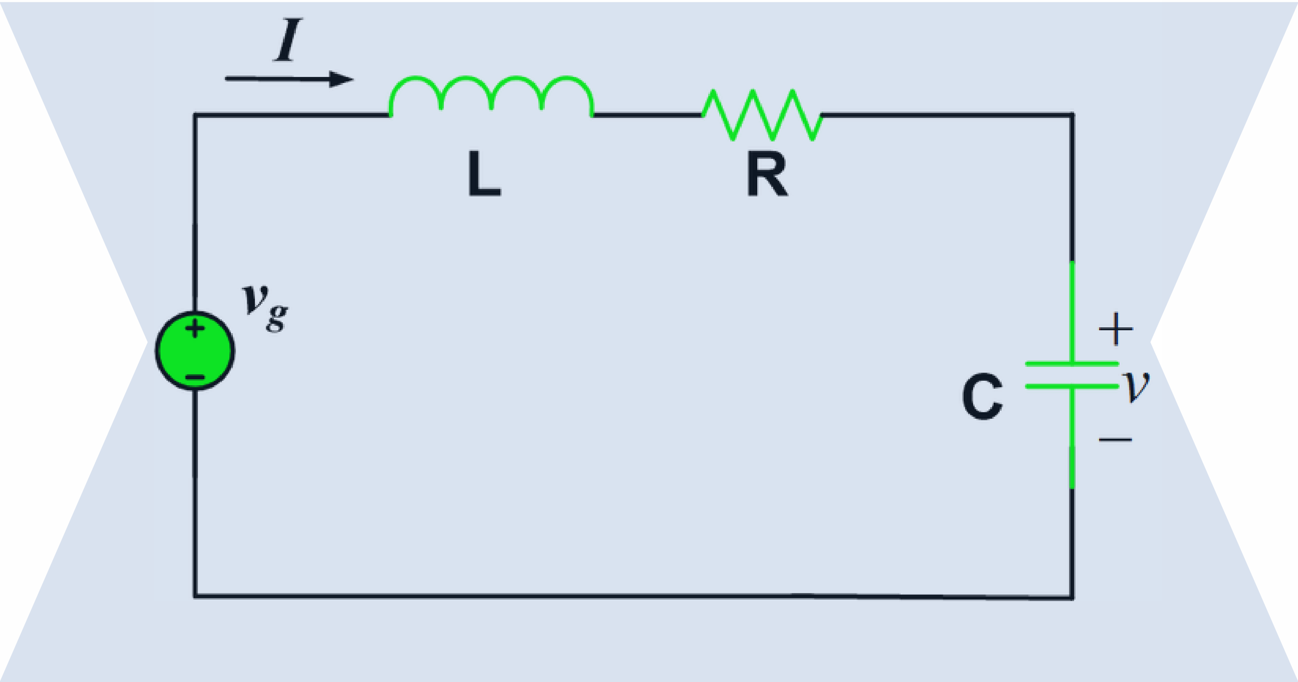
Electrical Engineering

GANDHI ACADEMY OF TECHNOLOGY AND ENGINEERING

CIRCUIT AND NETWORK THEORY



For 3rd Semester Electrical (Diploma)

Prepared by Susanta Kumar Sahu

As per SCTE&VT,Odisha Syllabus

Syllabus

1. **MAGNETIC CIRCUITS**
   - 1 . 1 Introduction
   - 1 . 2 Magnetizing force, Intensity, MMF, flux and their relations
   - 1 . 3 Permeability, reluctance and permeance
   - 1 . 4 Analogy between electric and Magnetic Circuits
   - 1 . 5 B-H Curve
   - 1 . 6 Series & parallel magnetic circuit.
   - 1 . 7 Hysteresis loop

2. **COUPLED CIRCUITS:**
   - 2 . 1 Self Inductance and Mutual Inductance
   - 2 . 2 Conductively coupled circuit and mutual impedance
   - 2 . 3 Dot convention
   - 2 . 4 Coefficient of coupling
   - 2 . 5 Series and parallel connection of coupled inductors.
   - 2 . 6 Solve numerical problems

3. **CIRCUIT ELEMENTS AND ANALYSIS:**
   - 3 . 1 Active, Passive, Unilateral & bilateral, Linear & Non linear elements
   - 3 . 2 Mesh Analysis, Mesh Equations by inspection
   - 3 . 3 Super mesh Analysis
   - 3 . 4 Nodal Analysis, Nodal Equations by inspection
   - 3 . 5 Super node Analysis.
   - 3 . 6 Source Transformation Technique
   - 3 . 7 Solve numerical problems (With Independent Sources Only)

4. **NETWORK THEOREMS:**
   - 4.1 Star to delta and delta to star transformation
   - 4.2 Super position Theorem
   - 4.3 Thevenin's Theorem
   - 4.4 Norton's Theorem
   - 4.5 Maximum power Transfer Theorem.
   - 4.6 Solve numerical problems (With Independent Sources Only)

5. **AC CIRCUIT AND RESONANCE:**
   - 5.1 A.C. through R-L, R-C & R-L-C Circuit
   - 5.2 Solution of problems of A.C. through R-L, R-C & R-L-C series Circuit by complex algebra method.
   - 5.3 Solution of problems of A.C. through R-L, R-C & R-L-C parallel & Composite Circuits

5.4 Power factor & power triangle.
5.5 Deduce expression for active, reactive, apparent power.
5.6 Derive the resonant frequency of series resonance and parallel resonance circuit
5.7 Define Bandwidth, Selectivity & Q-factor in series circuit.
5.8 Solve numerical problems

6. **POLYPHASE CIRCUIT**

6.1 Concept of poly-phase system and phase sequence
6.2 Relation between phase and line quantities in star & delta connection
6.3 Power equation in 3-phase balanced circuit.
6.4 Solve numerical problems
6.5 Measurement of 3-phase power by two wattmeter method.
6.6 Solve numerical problems.

7. **TRANSIENTS:**

7.1 Steady state & transient state response.
7.2 Response to R-L, R-C & RLC circuit under DC condition.
7.3 Solve numerical problems

8. **TWO-PORT NETWORK:**

8.1 Open circuit impedance (z) parameters
8.2 Short circuit admittance (y) parameters
8.3 Transmission (ABCD) parameters
8.4 Hybrid (h) parameters.
8.5 Inter relationships of different parameters.
8.6 T and π representation.
8.7 Solve numerical problems

9. **FILTERS:**

9.1 Define filter
9.2 Classification of pass Band, stop Band and cut-off frequency.
9.3 Classification of filters.
9.4 Constant – K low pass filter.
9.5 Constant – K high pass filter.
9.6 Constant – K Band pass filter.
9.7 Constant – K Band elimination filter.
9.8 Solve Numerical problems

CHAPTER-01 || MAGNETIC CIRCUITS

The closed path followed by magnetic flux is called a ***magnetic circuit.***

1. **Magnetic field:** The region around a magnet where its poles exhibit a force of attraction or repulsion is called *magnetic field*.
2. **Magnetic flux (ϕ):** The amount of magnetic lines of force set-up in a magnetic circuit is called *magnetic flux*. Its unit is weber (Wb). It is analogous to *electric current I* in electric circuit.
3. *The* **magnetic flux density** *at a point is the flux per unit area at right angles to the flux at that point*.

   It is, generally, represented by letter 'B'. Its unit is Wb/m^2 or Tesla, i.e.,

   $$B = \frac{\phi}{A}\ \text{Wb / m}^2 \quad \text{or} \quad \text{T } (1 \text{ Wb/m}^2 = 1 \times 10^4 \text{ Wb/cm}^2)$$

4. **Permeability:** *The ability of a material to conduct magnetic lines of force through it is called the* **permeability** *of that material.*

   It is generally represented by μ (*mu*, a Greek letter). The greater the permeability of a material, the greater is its conductivity for the magnetic lines of force and *vice-versa*. The permeability of air or vacuum is the poorest and is represented as μ_0 (where $\mu_0 = 4\pi \times 10^{-7}$ H/m).

5. **Magnetic field intensity:** The force acting on a unit north pole (1 Wb) when placed at a point in the magnetic field is called the magnetic intensity of the field at that point. It is denoted by *H*. In magnetic circuits, it is defined as mmf per unit length of the magnetic path. It is denoted by *H*, mathematically,

   $$H = \frac{\text{m.m.f}}{\text{length of magnetic path}} = \frac{NI}{l}\ \text{AT / m}$$

6. **Magnetomotive force (mmf):** The magnetic pressure which sets-up or tends to set-up magnetic flux in a magnetic circuit is called *magnetomotive force*. As per work law it may be defined as under:

   The work done in moving a unit magnetic pole (1 Wb) once round the magnetic circuit is called *magnetomotive force*. In general

   $$\text{mmf} = NI \text{ ampere-turns (or AT)}$$

   It is analogous to *emf* in an electric circuit.
7. **Reluctance (S):** The opposition offered to the magnetic flux by a magnetic circuit is called its *reluctance*.

   It depends upon length (l), area of cross-section (a) and permeability ($\mu = \mu_0 \mu_r$) of the material that makes up the magnetic circuit. It is measured in AT/Wb.

$$\text{Reluctance, } S = \frac{l}{a\,\mu_0\,\mu_r}$$

It is analogous to *resistance* in an electric circuit.

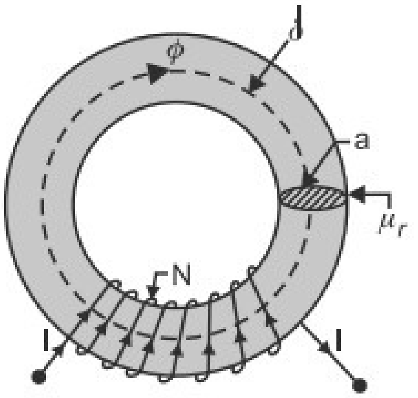
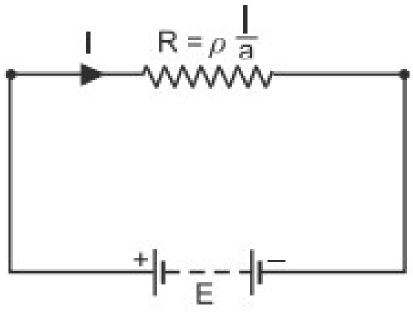
8. **Permeance:** It is a measure of the ease with which flux can be set-up in the material. It is just reciprocal of reluctance of the material and is measured in *Wb/AT* or *henry*.

$$\text{Permeance} = \frac{1}{\text{reluctance}} = \frac{a\,\mu_0\,\mu_r}{l} \text{ Wb/AT} \quad \text{or} \quad \text{H}$$

It is analogous to *conductance* in an electric circuit.

9. **Reluctivity:** It is specific reluctance and analogous to *resistivity* in electric circuit.

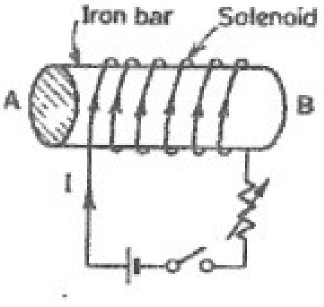
Comparison between Magnetic and Electric Circuits

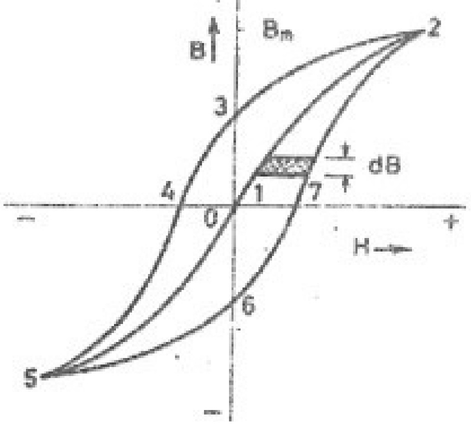
Magnetic Circuits	***Electrical Circuits***
 Magnetic circuit	 Electric circuit
Similarities	
1. The closed path for magnetic flux is called magnetic circuit. 2. Flux = mmf/reluctance 3. Flux, ϕ in Wb 4. mmf in AT 5. Reluctance, $S = \frac{l}{a\mu} = \frac{l}{a\mu_0\,\mu_r}$ AT/Wb 6. Permeance = 1/reluctance 7. Permeability, μ 8. Reluctivity 9. Flux density, $B = \frac{\phi}{a}$ Wb/m^2 10. Magnetic intensity, $H = NI/l$	1. The closed path for electric current is called electric circuit. 2. Current = emf/resistance 3. Current, I in ampere 4. emf in V 5. Resistance, $R = \rho \frac{l}{a} \Omega$ or $R = \frac{1}{\sigma} \frac{l}{a} \Omega$ 6. Conductance = 1/resistance 7. Conductivity, $\sigma = 1/\rho$ 8. Resistivity 9. Current density, $J = \frac{I}{a}$ A/m^2 10. Electric intensity, $E = V/d$

B.H. Curve :

Place a piece of an unmagnetised iron bar AB within the field of a solenoid to magnetise it. The field H produced by the solenoid, is called magnetising field, whose value can be altered (increased or decreased) by changing (increasing or decreasing) the current through the solenoid. If we increase slowly the value of magnetic field (H) from zero to maximum value, the value of flux density (B) varies

along 1 to 2 as shown in the figure and the magnetic materials (i.e iron bar) finally attains the maximum value of flux density (Bm) at point 2 and thus becomes magnetically saturated.





Now if value of H is decreased slowly (by decreasing the current in the solenoid) the corresponding value of flux density (B) does not decreases along 2-1 but decreases some what less rapidly along 2 to 3. Consequently during the reversal of magnetization, the value of B is not zero, but is '13' at H= 0. In other wards, during the period of removal of magnetization force (H), the iron bar is not completely demagnetized.

In order to demagnetise the iron bar completely, we have to supply the demagnetisastion force (H) in the opposite direction (i.e. by reserving the direction of current in the solenoid). The value of B is reduced to zero at point 4, when H='14'. This value of H required to clear off the residual magnetisation, is known as coercive force i.e. the tenacity with which the material holds to its magnetism.

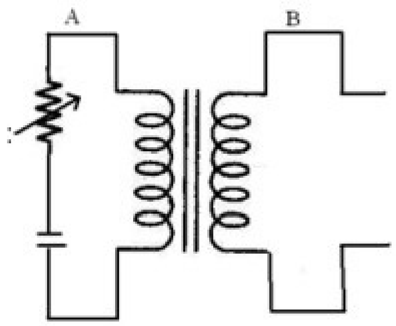
If after obtaining zero value of magnetism, the value of H is made more negative, the iron bar again reaches, finally a state of magnetic saturation at the point 5, which represents negative saturation. Now if the value of H is increased from negative saturation (= '45') to positive saturation (= '12') a curve '5,6,7,2' is obtained. The closed loop "2,3,4,5,6,7,2" thus represents one complete cycle of magnetisation and is known as **hysteresis loop**

CHAPTER-02 || COUPLED CIRCUITS

Inductance :→ It is defined as the property of the substance which opposes any change in Current & flux. Unit :→ Henry

Self-Inductance :→ It is defined as the property of the coil due to which it opposes any change (increase or decrease) of current or flux through it.

Mutual Inductance :→



It is defined as the emf induced in coil ‘B’ due to change of current in coil ‘A’

Co-efficient of Mutual Inductance (M) :

Thus, the fraction of magnetic flux produced by the current in one coil that links with the other is known as co-efficient of coupling (k) between the two coils.

Coefficient of mutual inductance between the two coils is defined as the weber-turns in one coil due to one ampere current in the other.

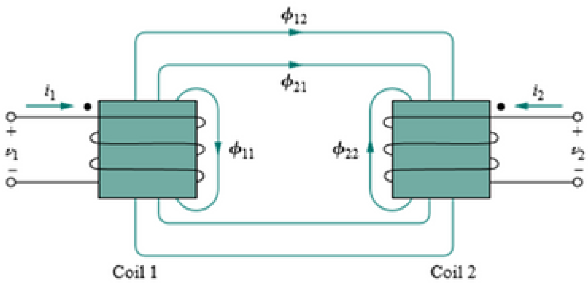
$$\boxed{\Rightarrow K = \sqrt{\frac{M.}{L_1.L_2}}}$$

Where ‘K’ is known as the co-efficient of coupling.

Co-efficient of coupling is defined as the ratio of mutual inductance between two coils to the square root of their self- inductances.

Dot convention:

- Required to determine polarity of “mutual” induced voltage.
- A dot is placed in the circuit at one end of each of the two magnetically coupled coils to indicate the direction of the magneticflux if current enters that dotted terminal of the coil



• Dot convention is stated as follows:

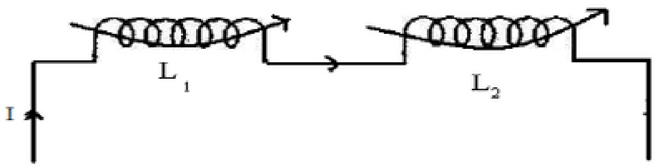
If a current ENTERS the dotted terminal of one coil, the reference polarity of the mutual voltage in the second coil is POSITIVE at the dotted terminal of the second coil.

• Conversely, Dot convention may also be stated as follow:

If a current LEAVES the dotted terminal of one coil, the reference polarity of the mutual voltage in the second coil is NEGATIVE at the dotted terminal of the second coil.

Inductance in series in Aiding

Inductances In Series (Additive) :→



Fluxes are in the same durection

Let M = Co-efficient of mutual inductance
L_1 = Co-efficient of self-inductance of first coil.
L_2 = Co-efficient of self-inductance of second coil.

EMF induced in first coil due to self-inductance

$$e_{L_1} = -L_1 \frac{dI}{dt}$$

Mutually induced emf in first coil

$$e_{M_1} = -M \frac{dI}{dt}$$

EMF induced in second coil due to self induction

$$e_{L_2} = -L_2 \frac{dI}{dt}$$

Mutually induced emf in second coil

$$e_{M_2} = -M \frac{dI}{dt}$$

Total induced emf

$$e = e_{L_1} + e_{L_2} + e_{M_1} + e_{M_2}$$

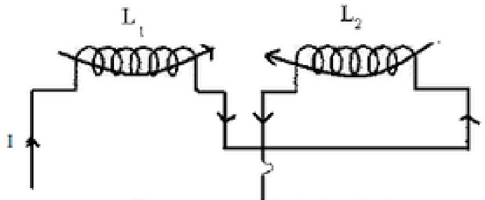
If ‘L’ is the equivalent inductance, then

$$-L\frac{dI}{dt} = -L_1\frac{dI}{dt} - M\frac{dI}{dt} - L_2\frac{dI}{dt} - M\frac{dI}{dt}$$

$$\Rightarrow -L\frac{dI}{dt} = -\frac{dI}{dt}(L_1 - L_2 - 2M)$$

$$\boxed{\Rightarrow L = L_1 + L_2 + 2M}$$

Inductance in series in Opposing



(Fluxes are opposite in direction)

Let M = Co-efficient of mutual inductance

L_1 = Co-efficient of self-inductance of first coil

L_2 -= Co-efficient of self-inductance of second coil

Emf induced in first coil due to self induction,

$$e_{L_1} = -L_1\frac{dI}{dt}$$

Mutually induced emf in first coil

$$e_{M_1} = -\left(-M\frac{dI}{dt}\right) = M\frac{dI}{dt}$$

Emf induced in second coil due to self-induction

$$e_{L_2} = -L_2\frac{dI}{dt}$$

Mutually induced emf in second coil

$$e_{M_2} = -\left(-M\frac{dI}{dt}\right) = M\frac{dI}{dt}$$

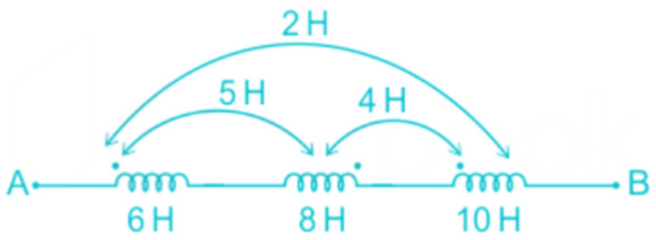
Total induced emf

$$e = e_{L_1} + e_{L_2} + e_{M_1} + e_{M_2}$$

Then $-L\frac{dI}{dt} = -L_1\frac{dI}{dt} - L_2\frac{dI}{dt} + M\frac{dI}{dt} + M\frac{dI}{dt}$

$$\Rightarrow -L\frac{dI}{dt} = -\frac{dI}{dt}(L_1 + L_2 - 2M) \qquad \Rightarrow L = L_1 + L_2 - 2M$$

Q. All the Three inductors are perfeclty coupled as shown below, the value of total inductance (in Henry) across the terminal AB is



Solution:

$$L_{eq} = 6 + 8 + 10 - (2 \times 5) - (2 \times 4) + (2 \times 2)$$

$$L_{eq} = 6 + 8 + 10 - 10 - 8 + 4$$

$$L_{eq} = 10 \text{ H}$$

CHAPTER-03 || CIRCUIT ELEMENTS AND ANALYSIS:

ACTIVE AND PASSIVE ELEMENT:

An active element has capability to generating energy while passive elements have not.

Ex: Active Element: Generators, Batteries, And Amplifiers. Passive Element: Resistor, Inductor, capacitor.

BILATERAL AND UNILATERAL ELEMENT:

If the magnitude of current passing through the element is affected due to change in the polarity of the applied voltage, the element is called unilateral element. And if the current magnitude remains same, it is called as bilateral element.

Ex: Unilateral Element: - Diodes, Transistors. Bilateral Element: - Resistor, Inductor, Capacitor

LINEAR AND NON-LINEAR ELEMENTS:

A linear element shows linear characteristics of voltage Vs current. Resistors, Inductor, Capacitor are linear elements and their property does not change in applied voltage on circuit current.
For non-linear elements the current passing through it does not change linearly with the time as change in

applied
voltage at a particular frequency.
Ex: Semiconductor devices

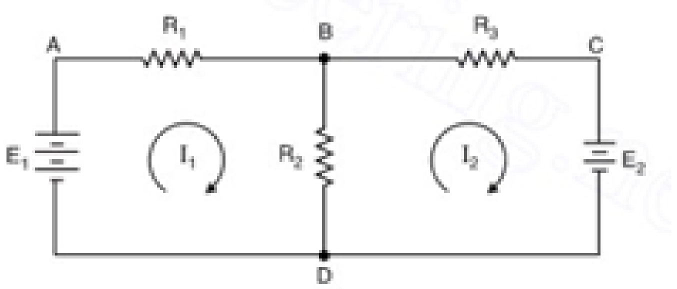
NODE ANALYSIS & MESH ANALYSIS

Two methods one Node analysis and the other mesh analysis are used to analyse a circuit depending on the arrangement and types of elements in the circuit. Nodal analysis is based on Kirchhoff's Current Law (KCL) and Mesh analysis is based on Kirchhoff's Voltage Law (KVL).

SUPERMESH

When a current source is common to two meshes we use the concept of super mesh to analysis the circuit using mesh current method. A super mesh is a larger mesh created from two meshes that have a current source as common element. A super mesh encloses more than one mesh for each common current source between two meshes, the number of meshes reduce by one, thus reading the number of mesh

Mesh Analysis:



Now applying Kirchhoff's voltage law to fig we get,

Mesh ABDA:

$$-I_1R_1-(I_1-I_2)R_2+E_1=0$$

Or, $I_1(R_1+R_2)-I_2R_2=E_1$ ________________(1)

Mesh BCDB:

$$-I_2R_3-(I_2-I_1)R_2-E_2=0$$

Or, $I_2(R_2+R_3)-I_1R_2=-E_2$ ________________(2)

For example, let $R_1=R_2=1, R_3=2$ & $E_1=5, E_2=10$, then putting the value in the above two equations, we get,

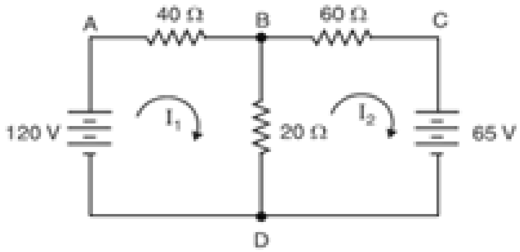
$2I_1-I_2=5$ .... (3)

$-I_1+3I_2=-10$............. (4)

Now solving the two equations, we get,

$I_1=1A$ & $I_2=-3A$ (Opposite to assume direction i.e., anticlockwise direction)

Q. In the network shown in below, find the magnitude and direction of each branch current by mesh analysis method?



Ans:

by mesh analysis:

Let us assign mesh currents $I_1 \& I_2$ to meshes ABDA and BCDB as shown in above. Now applying KVL to mesh **ABDA**, we get,

$$-40I_1 - 20(I_1 - I_2) + 120 = 0$$

Or, $60I_1 - 20I_2 = 120$.......................(1)

Applying KVL to mesh **BCDB,**

$$-60I_2 - 20(I_2 - I_1) - 65 = 0$$

Or, $-20I_1 + 80I_2 = -65$..................... (2)

Now solving the two equations (1) & (2), we get,

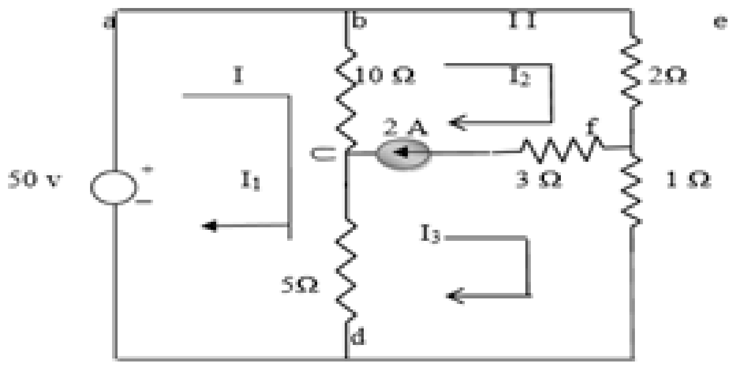
$I_1 = 1.88A$ & $I_2 = -0.34A$(Current actually flows in anticlockwise).

Current in branch **DAB** = $I_1 = 1.88A$

Current in branch **DCB**= $I_2 = 0.34A$

Current in branch **BD**= $I_1 + I_2 = 2.227A$

Q. Determine the current in the 5Ω resistor in the network given in fig. below



Solution:

From the mesh-1, i.e. abcda, we get,

$$50 = 10(I_1 - I_2) + 5(I_1 - I_3)$$

Or, $15I_1 - 10I_2 - 5I_3 = 50$ ....(1)

From the mesh-2 & 3, we can form a super mesh

$$0 = 10(I_2 - I_1) + 2I_2 + I_3 + 5(I_3 - I_1)$$

Or, $-15I_1 + 12I_2 + 6I_3 = 0$ ....(2)

The current source is equal to the difference between II and III mesh currents

i.e.

$I_2 - I_3 = 2A$ .... (3)

Now solving the three eq. (1), (2), (3), we get,

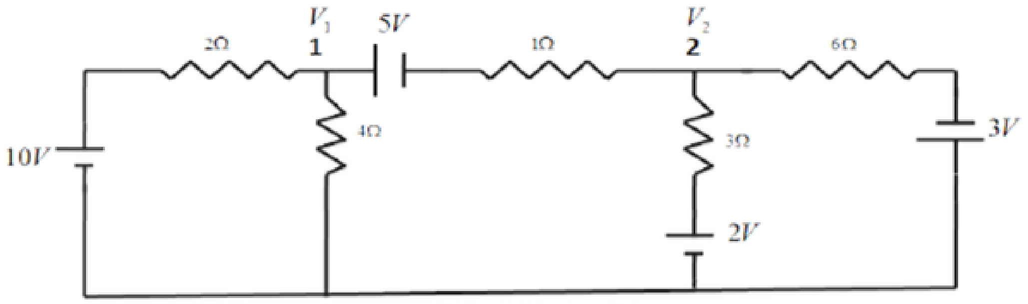
$I_1 = 19.99A, I_2 = 17.33A$ & $I_3 = 15.33A$

The current in 5Ω resistor $I_1 - I_3 = 19.99 - 15.33 = 4.66A$

Nodal Analysis:

- In nodal analysis it is essential to compute all the branch currents.
- In an N node circuit, one of the nodes is chosen as the reference node or datum node.
- Each node in a circuit can be assigned a number or letter.
- The node voltage is the voltage of a given node with respect to one particular node, called reference node, which will assume at zero potential.
- In writing the current expression, the node potential is always taken as higher potential, than the other voltages appearing in the network

Q. Write down the nodal voltages for the given circuit



3 (Reff. Node=0V)

Solution: By applying nodal analysis,

Let us first assign a voltage to each node. There are 3 no of nodes that are 1, 2 & 3 and have the voltages V_1, V_2, V_3.

As node (3) is common to all the nodes, so take node-3 as the reference node i.e. 0V.

Assuming that at node-1, all the currents are leaving. So applying KCL to node-1, we get,

$$\frac{V_1 - 10}{2} + \frac{V_1 - 0}{4} + \frac{V_1 - 5 - V_2}{1} = 0$$

Or, $V_1\left[\frac{1}{2} + \frac{1}{4} + 1\right] - V_2 = 10$

Or, $1.75V_1 - V_2 = 10$ .... (1)

At node-2, assuming all the currents are leaving expect the current from current source, we get,

$$\frac{V_2 + 5 - V_1}{1} + \frac{V_2 - 2}{3} + \frac{V_2 + 3}{6} = 0$$

Or, $-V_1 + V_2\left[1 + \frac{1}{3} + \frac{1}{6}\right] = -5 + \frac{2}{3} - \frac{1}{2}$

Or, $-V_1 + 1.5V_2 = -4.833$ .... (2)

Solving eqn 1 and 2 we will get V1=6.25V and V2=0.95V

Supernode:

Suppose any of the branches in the network has a voltage source; then it is slightly difficult to apply nodal analysis. One way to overcome this difficulty is to apply the supernode technique. *In this method, the two adjacent nodes that are connected by a voltage source are reduced to a single node and then the equations are formed by applying Kirchhoff's current law as usual.* This is explained with the help of Fig.

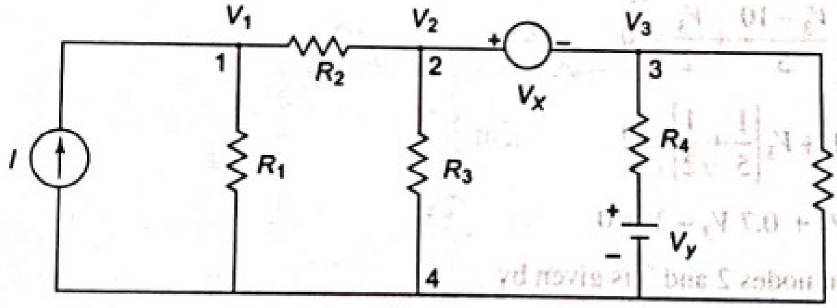


Fig.

It is clear from Fig. 2.44, that the node 4 is the reference node. Applying Kirchhoff's current law at the node 1, we get

$$I=\frac{V_1}{R_1}+\frac{V_1-V_2}{R_2}$$

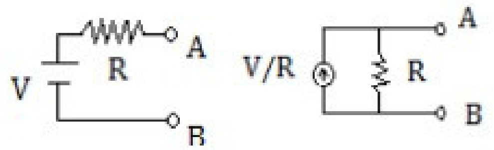
Due to the presence of voltage source V_x in between nodes 2 and 3, it is slightly difficult to find out the current. The supernode technique can be conveniently applied in this case.

Accordingly, we can write the combined equation for nodes 2 and 3 as under.

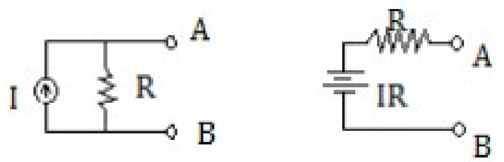
$$\frac{V_2-V_1}{R_2}+\frac{V_2}{R_3}+\frac{V_3-V_y}{R_4}+\frac{V_3}{R_5}=0$$

SOURCE TRANSFORMATIONS

In the circuit analysis, a circuit with either voltage source or current sources is preferred. Sometimes a circuit may have both i.e. voltage source & current source. In that case it is convenient to transform voltage source to equivalent current source and current source to equivalent voltage source .



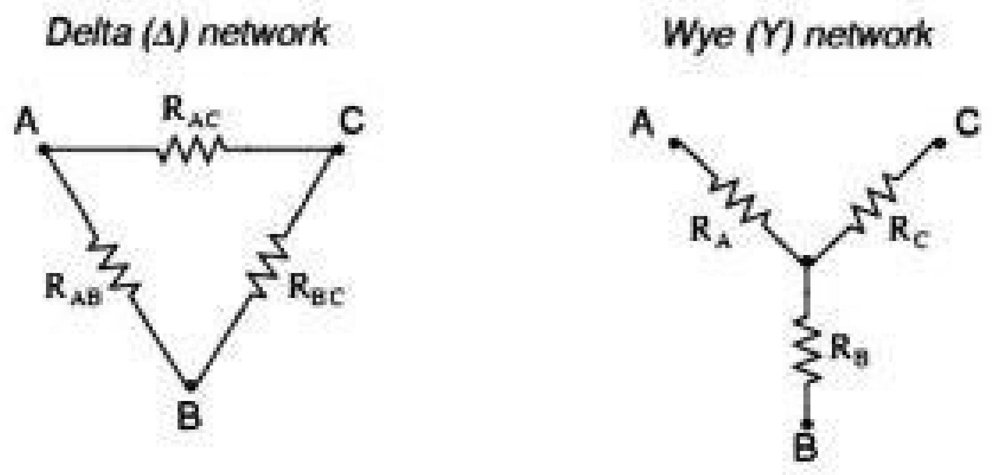
(Transformation of Voltage source to an equivalent current source)



(current source to an equivalent voltage source)

CHAPTER-04 || NETWORK THEOREMS:

Star to delta and delta to star transformation



To convert a Delta (Δ) to a Wye (Y)

$$R_A = \frac{R_{AB}\,R_{AC}}{R_{AB} + R_{AC} + R_{BC}}$$

$$R_B = \frac{R_{AB}\,R_{BC}}{R_{AB} + R_{AC} + R_{BC}}$$

$$R_C = \frac{R_{AC}\,R_{BC}}{R_{AB} + R_{AC} + R_{BC}}$$

To convert a Wye (Y) to a Delta (Δ)

$$R_{AB} = \frac{R_A R_B + R_A R_C + R_B R_C}{R_C}$$

$$R_{BC} = \frac{R_A R_B + R_A R_C + R_B R_C}{R_A}$$

$$R_{AC} = \frac{R_A R_B + R_A R_C + R_B R_C}{R_B}$$

Super position Theorem

In a linear bilateral network containing two or more independent sources, the voltage across or current in any branch is algebraic sum of individual voltages or currents produced by each independent sources acting separately with all the independent sources set equal to zero.

Explanation

In Fig. a to apply superposition theorem, let us first take the source V_1 alone at first replacing V_2 by short circuit Fig.a-1

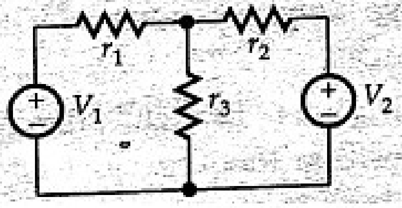


Fig. a

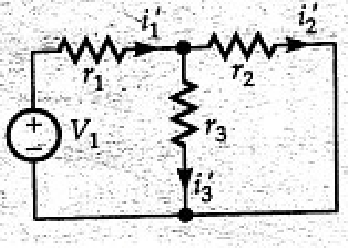


Fig.a-1

$$\text{Here,}\quad i_1' = \frac{V_1}{\dfrac{r_2 r_3}{r_2 + r_3} + r_1}$$

$$i_2' = i_1' \frac{r_3}{r_2 + r_3}$$

$$\text{and}\quad i_3' = i_1' - i_2'$$

Next, removing V_1 by short circuit, let the circuit be energized by V_2 only Fig.a-2

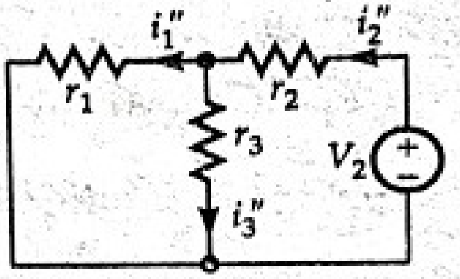


Fig.a-2

$$\text{Here,}\quad i_2'' = \frac{V_2}{\dfrac{r_1 r_3}{r_1 + r_3} + r_2} \quad \text{and} \quad i_1'' = i_2'' \frac{r_3}{r_1 + r_3}$$

$$\text{Also,}\quad i_3'' = i_2'' - i_1''.$$

As per Superposition theorem,

$$i_3 = i_3' + i_3''$$

$$i_2 = i_2' - i_2''$$

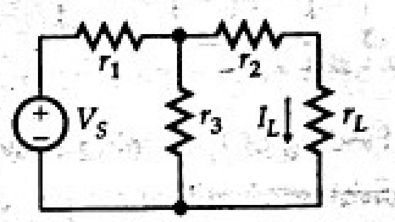
$$i_1 = i_1' - i_1''$$

Thevenin’s Theorem

Thevenin’s theorem states that any linear active two terminal network containing resistance and voltage sources or current sources can be replaced by a single voltage sources Vth in series with single resistance Rth. The Thevenin equivalent voltage Vth is the open circuit voltage at the network terminal and the Thevenin resistance Rth. is the resistance between the network terminals when all the sources are replaced with their internal resistance.

Explanation

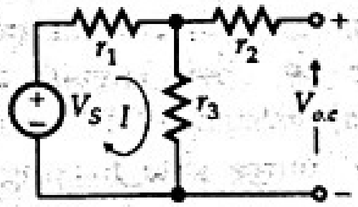
Let us consider a simple d.c. circuit as shown in Fig. (*a*). We are to find I_L by Thevenin's theorem.



(*a*) A Simple d.c. circuit

In order to find the equivalent voltage source, r_L is removed (Fig. (*b*)) and $V_{o.c}$ is calculated.

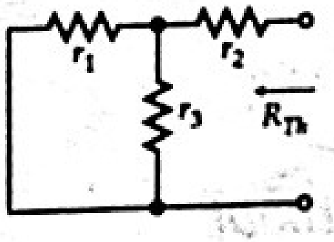
$$V_{o.c} = I\, r_3 = \frac{V_S}{r_1 + r_3} \cdot r_3$$



(*b*) Finding of V_{oc}

Next, to find the internal resistance of the network (*Thevenin's resistance* or *equivalent resistance*) in series with $V_{o.c}$, the voltage source is removed (deactivated) by a short circuit (as the source does not have any internal resistance) as shown in Fig. (*c*).

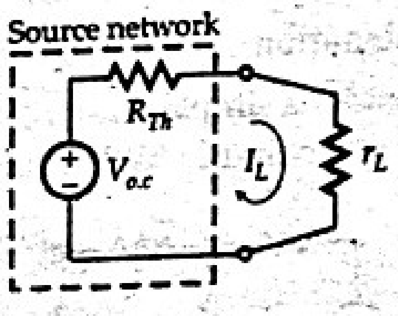
$$R_{Th} = r_2 + \frac{r_1 r_3}{r_1 + r_3}.$$



(*c*) Finding of R_{Th}

As per Thevenin's theorem, the equivalent circuit being Fig. (*d*),

$$I_L = \frac{V_{o.c}}{R_{Th} + r_L} \text{ A}$$



(*d*) Finding of I_L forming Thevenin's equivalent circuit

Norton's Theorem

Statement of Norton's Theorem

A linear active network consisting of independent and or dependant voltage and current sources and linear bilateral network elements can be replaced by an equivalent circuit consisting of a current source in parallel with a resistance, the current source being the short circuited current across the load terminal and the resistance being the internal resistance of the source network looking through the open circuited load terminals.

Explanation

In order to find the current through r_L, the load resistance (Fig. 3.3), by Norton's theorem, let us replace r_L by short circuit (Fig. 3.3(*a*)).

Obviously, $i = \dfrac{V_S}{r_1 + \dfrac{r_2 r_3}{r_2 + r_3}}$ and $i_{sc} = i \dfrac{r_3}{r_3 + r_2}$

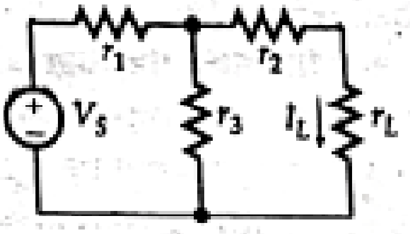


Fig. 3.3 A simple dc network

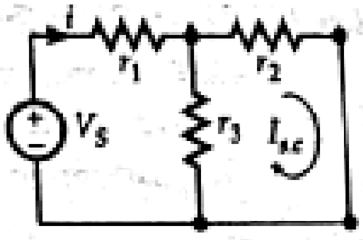


Fig. 3.3 (*a*) Finding of i_{sc}

Next, the short circuit is removed and the independent source is deactivated as done in Thevenin's theorem (Fig. 3.3(*b*)).

Here, $R_{int} = r_2 + \frac{r_1 r_3}{r_1 + r_3}$.

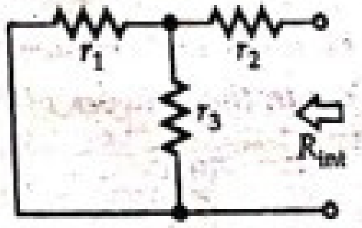


Fig. 3.3 (*b*) Finding of R_{Th} (or R_{int})

As per Norton's theorem, the equivalent source circuit would contain a current source in parallel to the internal resistance, the current source being the short circuited current across the shorted terminals of the load resistor (Fig. 3.3(*c*)).

Obviously, $I_L = i_{s.c} \frac{R_{int}}{R_{int} + r_L}$

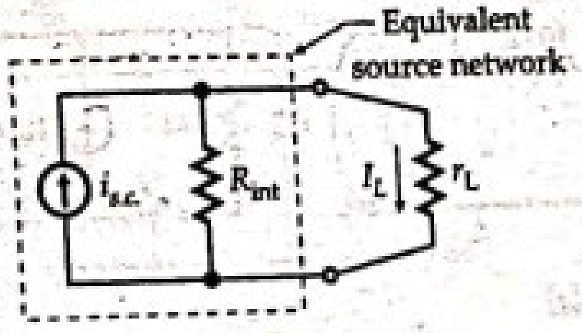
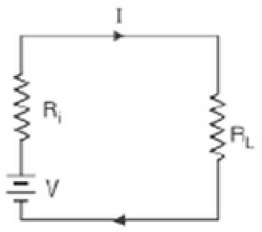


Fig. 3.3 (*c*) Norton's equivalent circuit

Maximum power Transfer Theorem.

In a linear bilateral network containing an independent voltage source in series with resistance R_S delivers maximum power to the load resistance R_L when $R_L=R_S$

Consider the voltage source V of internal resistance R_i delivering power to a load R_L. We shall prove that when $R_L = R_i$, the power delivered to R_L is maximum. Referring to the fig. below, we have,



Circuit current, $I = \dfrac{V}{R_L + R_i}$

Power delivered to load, $P = I^2 R_L = \left(\dfrac{V}{R_L + R_i}\right)^2 R_L$ .... (i)

For a given source, generated voltage V and internal resistance R_i are constant. Therefore, power delivered to the load depends upon R_L. In order to find the value of R_L for which the value of P is maximum, differentiate eq. (i) w.r.t. R_L and set the result equal to zero.

Thus, $$\frac{dP}{dR_L} = V^2\left[\frac{(R_L + R_i)^2 - 2R_L(R_L + R_i)}{(R_L + R_i)^4}\right] = 0$$

Or, $$(R_L + R_i)^2 - 2R_L(R_L + R_i) = 0$$

Or, $$(R_L + R_i)(R_L + R_i - 2R_L) = 0$$

Or, $$(R_L + R_i)(R_i - R_L) = 0$$

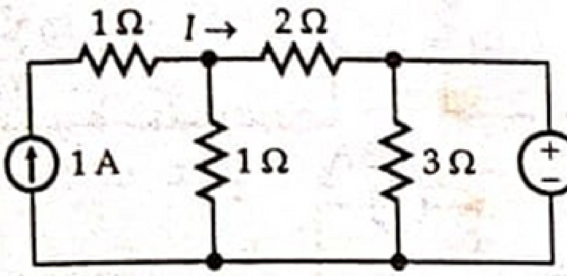
$(R_L + R_i)$ cannot be zero,

$$(R_i - R_L) = 0$$

$$R_L = R_i$$

$$\textit{Load resistance} = \textit{Internal resistance of the source}$$

Q. *Find I in the circuit shown in Fig.*

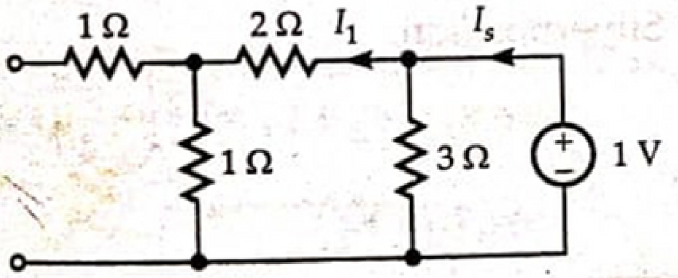


SOLUTION. Principle of Superposition is applied by taking 1 V source only at first

$$I_s = \frac{1\,\text{V}}{[(1+2)||3]\Omega} = (1/1.5)\ \text{A}$$

$$I_1 = I_s \frac{3}{3+2+1} = \frac{1}{1.5} \times \frac{3}{6} = \frac{1}{3}\ \text{A}$$

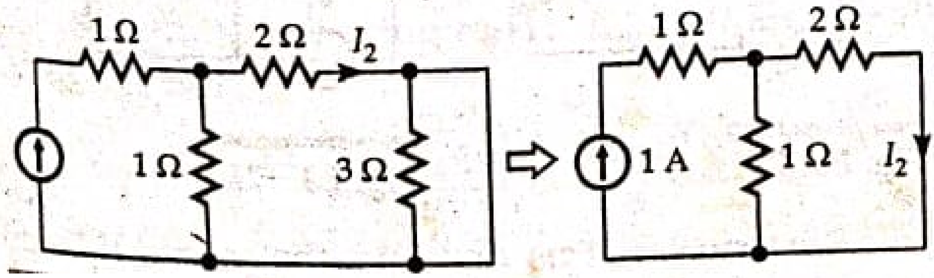
[by current division formula]



Next, let us assume the current source only

$$I_2 = 1 \times \frac{1}{1+2} = \frac{1}{3}\ \text{A}$$

[by current division formula]

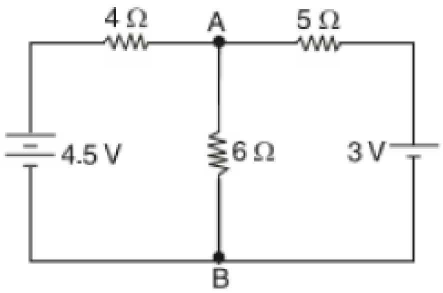


It may be observed that utilising the principle of Superposition, the net response can be obtained when both the sources (1 A and 1 V) are present. The current through 2 Ω resistor is obtained as

$$I = (I_1 - I_2) = \frac{1}{3} - \frac{1}{3} = 0$$

[I_1 and I_2 being directed reverse].

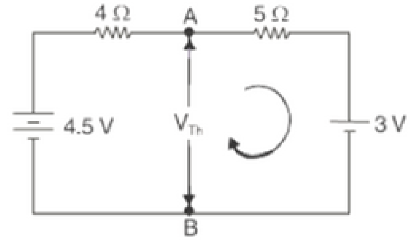
Q. *Using Thevenin's theorem, find the current in 6Ω resistor in Fig. (i) shown in below.*



(i)

Solution:

STEP-1: Finding V_{Th} :



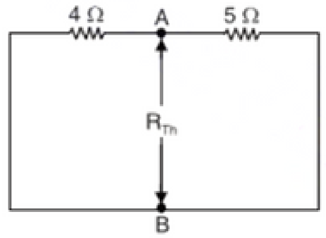
(ii)

Since the internal resistance of the battery is not given, it will assume that they are zero. Now remove the load resistance $R_L = 6\Omega$, Find the Thevenin's equivalent voltage $V_{Th} = V_{AB}$ across **AB** terminal.

Net e.m.f. in the circuit shown in Fig. (ii) is *4.5-3=1.5V* and total circuit resistance is 9Ω.

∴ Circuit current = 1.5/9 = 0.167A

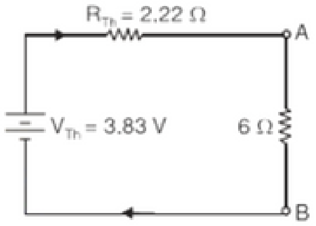
STEP-2: Finding R_{Th} :



R_{Th} = Resistance at terminals ***AB*** with load (i.e. 6Ω resistor) removed and battery replaced by a short as shown in Fig. above.

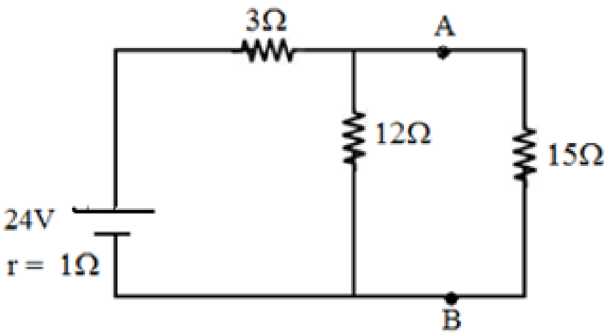
$$R_{Th} = \frac{4 \times 5}{4+5} = 2.22\Omega$$

STEP-3: Now draw the Thevenin's equivalent circuit to find the load current i.e. current in 6Ω resistor:

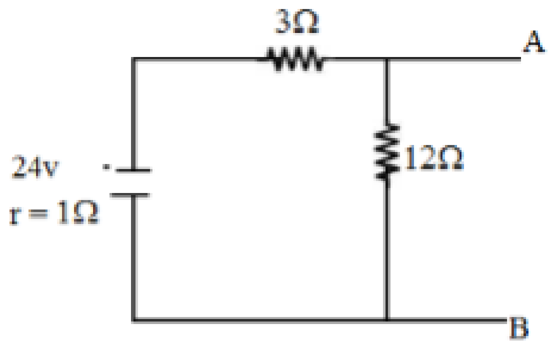


$\therefore$ Current in 6Ω resistor i.e. $I_L = I_6\Omega = \dfrac{V_{Th}}{R_{Th} + R_L} = \dfrac{3.83}{2.22+6} = 0.466A$

Q. Applying Thevenin's theorem find the current in 15Ω resistance shown in below Fig

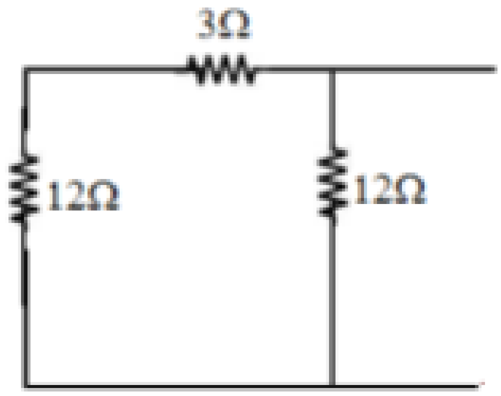


Solution:



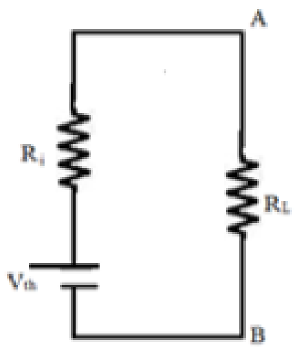
Remove 15Ω resistance and find the voltage across A and B i.e. $V_{Th} = V_{AB}$

$$V_{Th} = V_{AB} = \frac{24 \times 12}{12 + 3 + 1} = 18V$$



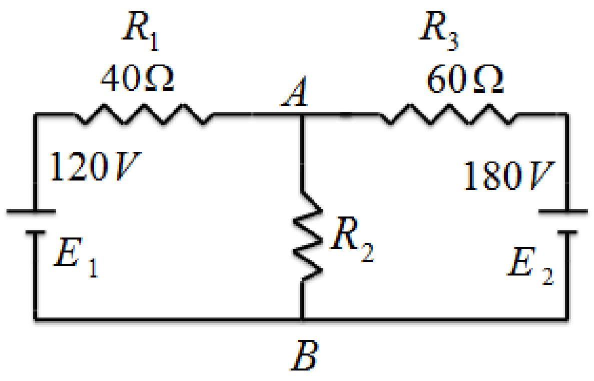
R_{Th} is calculated from the terminal A & B into the network. The 1Ω and 3Ω are in series and their equivalent is parallel to 12Ω resistance.

$$R_{Th} = \frac{4 \times 12}{16} = 3\Omega$$

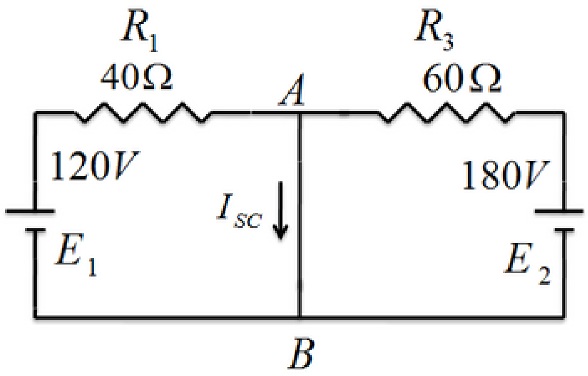


$\therefore$ Current in 15Ω resistor i.e. $I_L = I_{15}\Omega = \dfrac{V_{Th}}{R_{Th} + R_L} = \dfrac{18}{15 + 3} = 1A$

Q. *Using Norton's theorem find the current that would flow through the resistor R_2 when it takes the value of 12Ω, in the fig shown below.*



Remove the load resistance by making short circuit of terminal AB :

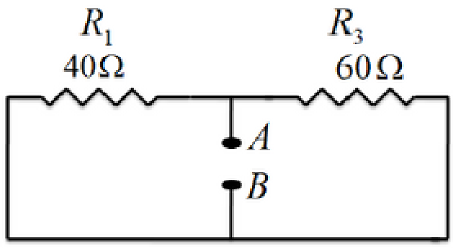


Finding the short circuit current Isc

First the current due to E_1 is$=\dfrac{120}{40}=3A$, and due to E_2 is $=\dfrac{180}{60}=3A$

Then $I_{SC}=3+3=6A$

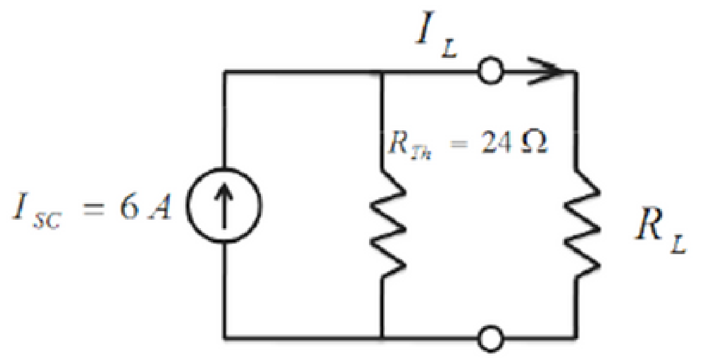
Finding the Norton's equivalent resistance i.e.equal to Thevenin's equivalent resistance $R_N = R_{Th}$:



It is calculated by open circuit load resistance and viewed from open circuit and into the network and all sources are taken to zero.

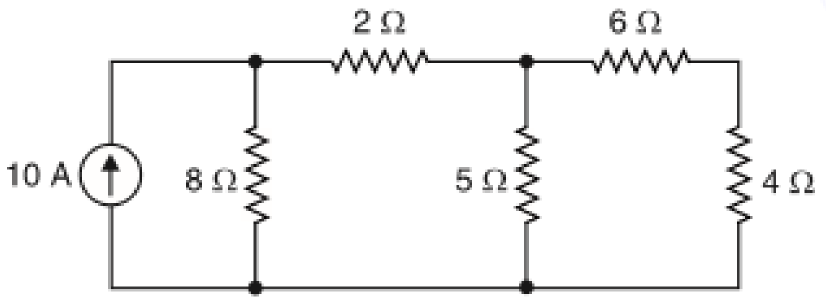
$$R_N = 40 \| 60 = \frac{40 \times 60}{40 + 60} = 24\Omega$$

Now draw the Norton's equivalent circuit to find the load current I_L :

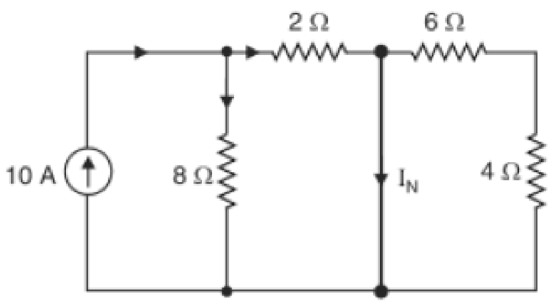


When $R_L = 12\Omega, I_L = I_{SC} \dfrac{R_{Th}}{R_{Th} + R_L} = \dfrac{6 \times 24}{24 + 12} = 4A$

Q. Using Norton's theorem, calculate the current in the 5Ω resistor in the circuit shown in fig. below



Remove the load resistance by making short circuit of terminal AB :

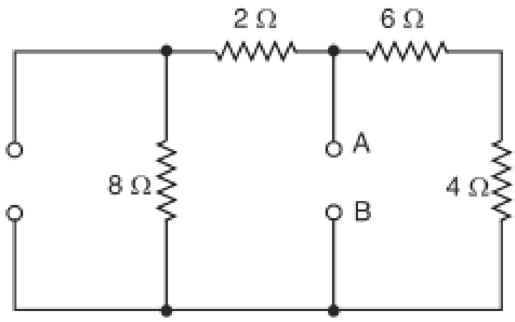


Finding the short circuit current $I_{SC} = I_N$:

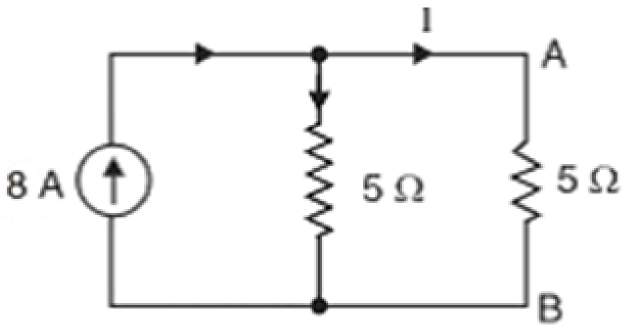
As the load terminals are short-circuited, and then the series combination of 6Ω and 4Ω are in same potential and is disconnected. Now current in 2Ω resistance is same to the Norton's equivalent current I_N and may be calculated by the current divider rule,

$$I_N = 10 \times \frac{8}{8+2} = 8A$$

Finding the Norton's equivalent resistance i.e. equal to Thevenin's equivalent resistance $R_N = R_{Th}$:

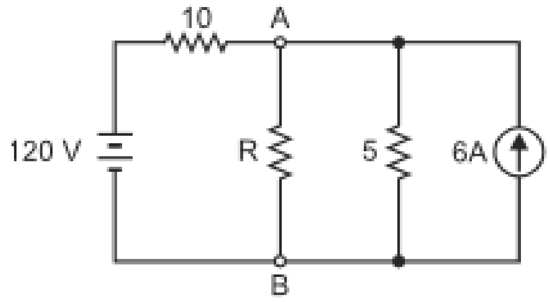


$$R_N = R_{Th} = R_{AB} = (8+2) \| (6+4) = \frac{10 \times 10}{10+10} = 5\Omega$$



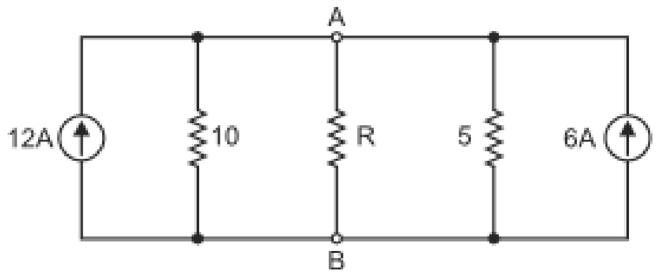
$$I_L = I = I_N \frac{R_{Th}}{R_{Th} + R_L} = \frac{8 \times 5}{5+5} = 4A$$

Q. Find the value of resistance R to have maximum power transfer in the circuit shown in Fig. Also obtain the amount of maximum power. All resistances are in ohms.

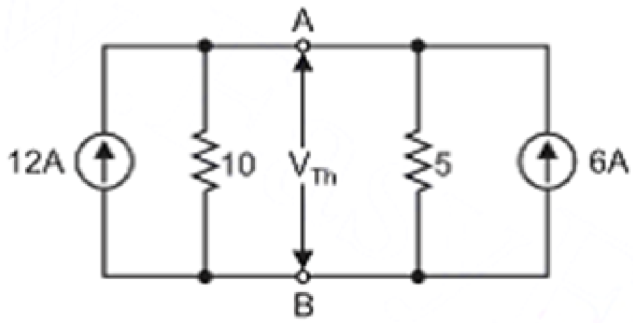


Solution:

To find the value of resistance **R**, We have to find the V_{Th} & R_{Th} at the load terminals i.e. ***(R)***. Now convert the voltage source 120V series with resistance 10Ω to equivalent current source of 120/10=12A in parallel with 10Ωresistance as shown in fig. below.



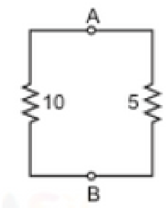
To find V_{Th} (i.e. voltage across the ***A*** and ***B*** terminals) applying the Nodal analysis,



$$\frac{V_{Th}}{10} + \frac{V_{Th}}{5} = 12 + 6$$

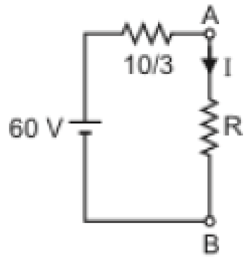
Or, $V_{Th} = \frac{180}{3} = 60V$

In order to find R_{Th} , remove ***R*** and replace the current sources by open circuit. So the $R_{Th} = R_{AB}$ is given as,



$$R_{Th} = R_{AB} = \frac{10 \times 5}{10+5} = \frac{10}{3} = 3.33\Omega$$

When R is connected to the terminals of Thevenin's equivalent circuit, the circuit becomes as shown in below fig.

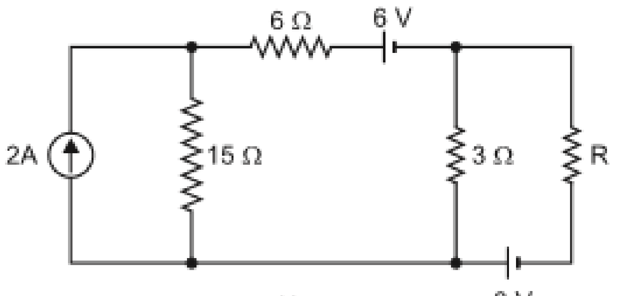


For maximum power transfer, the condition is

$$R = R_{Th} = \frac{10}{3}\Omega$$

Max. Power transferred, $P_{\max} = \frac{V_{Th}^2}{4R_L} K = \frac{(60)^2}{4 \times (10/3)} = 270W$

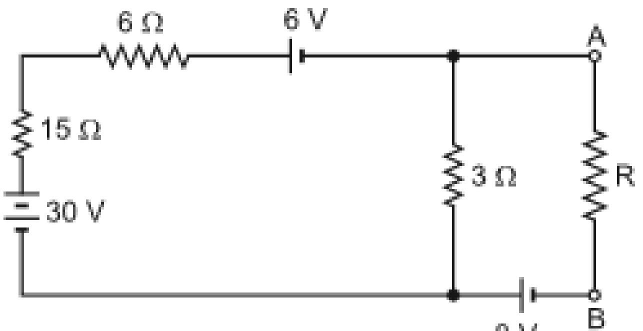
Q. Calculate the value of resistance R which will absorb maximum power from the circuit shown in Fig. (i) below



(i)

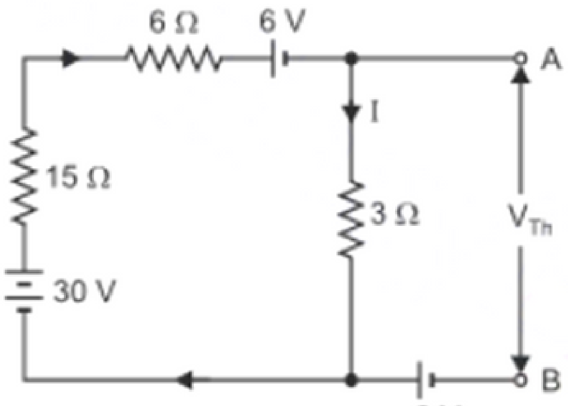
Solution:

To find the value of resistance ***R***, We have to find the V_{Th} & R_{Th} at the load terminals i.e. ***(R)***. Now convert the current source 2A parallel with resistance 15Ω to equivalent voltage source of $2 \times 15 = 30V$ in series with 15Ωresistance as shown in fig. (ii) below.



(ii)

To find V_{Th} (i.e. voltage across the ***A*** and ***B*** terminals) referring to the fig. (iii) below,



(iii)

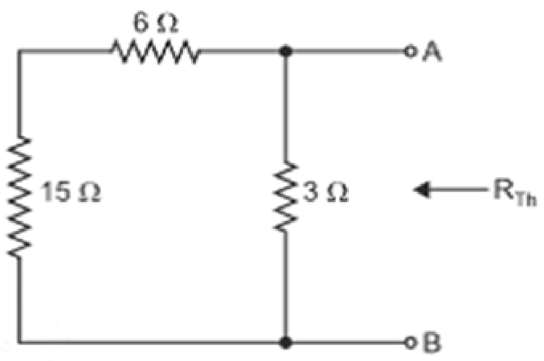
Current in the 3Ω resistor, $I = \dfrac{30-6}{15+6+3} = \dfrac{24}{24} = 1A$

Now applying KVL in the right side open loop, we get,

$$-V_{Th} + (3 \times 1) + 8 = 0$$

Or, $V_{Th} = 3 + 8 = 11V$

In order to find R_{Th} , remove ***R*** and replace the voltage sources by short circuit as shown in fig. (iv) below. So the $R_{Th} = R_{AB}$ is given as,



(*iv*)

$$R_{Th} = R_{AB} = (6+15)\Omega \| 3\Omega = \frac{21\times 3}{21+3} = \frac{21}{8}\Omega$$

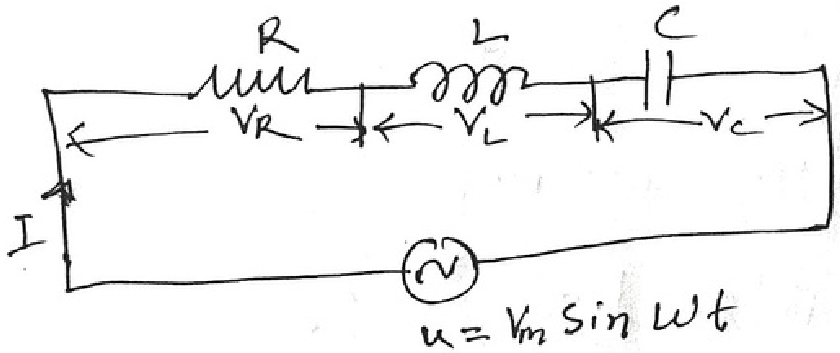
For maximum power transfer, the condition is

$$R = R_{Th} = \frac{21}{8}\Omega$$

Max. Power transferred, $P_{\max} = \dfrac{V_{Th}^2}{4R_L} = \dfrac{(11)^2}{4\times(21/8)} = 11.524W$

CHAPTER -05 || AC CIRCUIT AND RESONANCE:

R-L-C Series Circuit



$V_R = IR$

$V_L = IX_L$ and $X_L = 2\pi f L$

$V_C = IX_C$ and $X_C = \frac{1}{2\pi f C}$

$V = \sqrt{V_R^2 + (V_L - V_C)^2}$

Impedance $(Z) = \sqrt{R^2 + (X_L - X_C)^2}$

Current, $I = \frac{V}{Z}$

Power factor, $\cos\phi = \frac{R}{Z}$

Active Power, $P = VI\cos\phi$, watt

Reactive Power, $Q = VI\sin\phi$, VAR

Apparent Power, $S = VI$, VA

Q. The current in a circuit is given by (4.5 + j12) A when the applied voltage is (100 + j150) V. Determine (i) the magnitude of impedance and (ii) phase angle

Sol:

$$V = (100 + j50)V = 180.28\angle 56.31^\circ V$$

$$I = (4.5 + j12)A = 12.82\angle 69.44^\circ A$$

(i) $\therefore\ Z = \dfrac{V}{I} = \dfrac{180.28\angle 56.31^\circ}{12.82\angle 69.44^\circ} = 14.06\angle -13.13^\circ \Omega$

$\therefore\ Z = 14.06\Omega$

(ii) $\therefore$ Phase angle, $\Phi = 13.13^\circ lead$

Q2. In an R-L series circuit, $R = 10\Omega$ and $X_L = 8.66\Omega$. If current in the circuit is $(5 - j10)A$, find (i) the applied voltage (ii) power factor and (iii) active power and reactive power.

Sol:

$$Z = (R + jX_L) = (10 + j8.66)\Omega$$

$$= 13.23\angle 40.9^\circ \Omega$$

$$I = (5 - j10)A = 11.18\angle -63.43^\circ A$$

(i) Applied voltage, $V = IZ$

$= 11.18\angle -63.43^\circ \times 13.23\angle 40.9^\circ$

$= 148\angle -22.53^\circ V$

$V = 148 volts$

(ii) Phase angle, $\Phi = 63.43^\circ - 22.53^\circ = 40.9^\circ$

Power factor $= \cos\Phi = \cos 40.9^\circ = 0756 lag$

(iii) Complex VA, **S**=Phase voltage $\times$ Conjugate of phasor current

Or, $P + jQ = 148\angle -22.53^\circ \times 11.18\angle 63.43^\circ = 1654.64\angle 40.9^\circ VA$

$$= 1654.64(\cos 40.9^\circ + j\sin 40.9^\circ) = (1250.66 + j1083.36)VA$$

$\therefore$ Active power $P = 1250.66W$; Reactive power, $Q = 1083.36VAR$

Q.3. A coil of resistance $R = 12\Omega$ and inductive reactance $(X_L) = 25\Omega$ is connected in series with a capacitive reactance of $X_C = 41\Omega$. The combination is connected to a supply of 230V, 50Hz. Using phasor algebra, find (i) circuit impedance (ii) current and (iii) power consumed.

Sol:

(i) $Z = R + j(X_L - X_C)$

$= 12 + j(25 - 41) = (12 - j16)\Omega = 20\angle -53.13^\circ \Omega$

$\therefore$ $Z = 20\Omega$

(ii) Taking voltage as the reference phasor,

$V = (230 + j0)V = 230\angle 0^\circ volts$

$$\therefore \quad I = \frac{V}{Z} = \frac{230\angle 0^\circ}{20\angle -53.13^\circ} = 11.5\angle 53.13^\circ A$$

$\therefore$ $I = 11.5A$

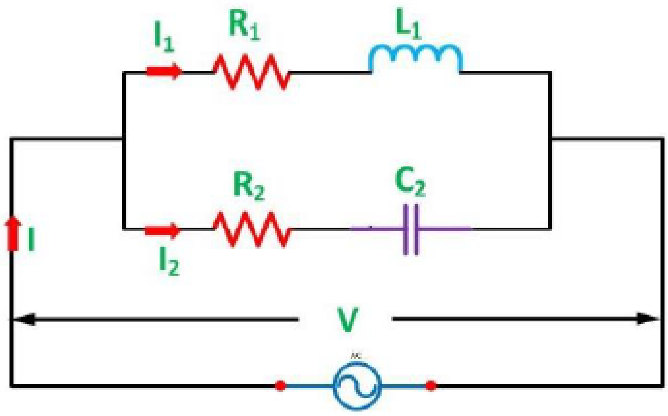
(iii) It is clear that current leads the voltage by 53.13°i.e. $\Phi = 53.13^\circ$

$\therefore$ Power factor $= \cos\Phi = \cos 53.13^\circ = 0.6 lead$

Power consumed, $P = VI\cos\Phi = 230 \times 11.5 \times 0.6 = 1587W$

Steps for Solving Circuit by Admittance Method

Consider a parallel AC circuit having resistance and capacitance connected in series and resistance and inductance also connected in series as shown in the figure below.



- **Step 1**
  - Draw the circuit as per the given problem.

- **Step 2**
  - Impedance of the branch 1 is given by $Z_1 = R_1 + jX_{L1}$
    Impedance of the branch 2 is given by $Z_2 = R_2 - jX_{C2}$

- **Step 3**
  - Find the equivalent impedance

$$Z_{equ} = \frac{Z_1 \times Z_2}{Z_1 + Z_2} = \frac{(R_1 + jX_{L1}) \times (R_2 - jX_{C2})}{R_1 + jX_{L1} + R_2 - jX_{C2}}$$

- **Step 4**
  - Find the Total Current 'I'

$$I = \frac{V}{Z_{equ}}$$

- **Step 5**
  - Find the branch Current

$$I_1 = \frac{I \times Z_2}{Z_1 + Z_2} \quad and \quad I_2 = \frac{I \times Z_1}{Z_1 + Z_2}$$

- **Step6**
  - Find the Power by $S = P + jQ \quad S = V\,I^*$
    Where I^* is the complex conjugate of I

The real part P indicates the Active or real or True Power and Q indicates the Reactive power.

Q. . Two impedances $Z_1 = (8 + j6)\Omega$ and $Z_2 = (3 - j4)\Omega$ are connected in parallel. If the total current of this combination is 25A, find the power taken by each impedance.

Sol:

$$Z_1 = (8 + j6)\Omega = 10\angle 36.87^\circ \Omega$$

$$Z_2 = (3 - j4)\Omega = 5\angle -53.13^\circ \Omega$$

$$Z_1 + Z_2 = (8 + J6) + (3 - J4) = (11 + J2)\Omega = 11.18\angle 10.3^\circ \Omega$$

$$I_1 = I\frac{Z_2}{Z_1 + Z_2} = 25 \times \frac{5\angle -53.13^\circ}{11.18\angle 10.3^\circ} = 11.18\angle -63.43^\circ A$$

$$I_2 = I\frac{Z_1}{Z_1 + Z_2} = 25 \times \frac{10\angle 36.87^\circ}{11.18\angle 10.3^\circ} = 22.36\angle 26.57^\circ A$$

Power taken by first branch $= I_1^2 R_1 = (11.18)^2 \times 8 = 1000W$

Power taken by first branch $= I_2^2 R_2 = (22.36)^2 \times 3 = 1500W$

Power factor & Power triangle:

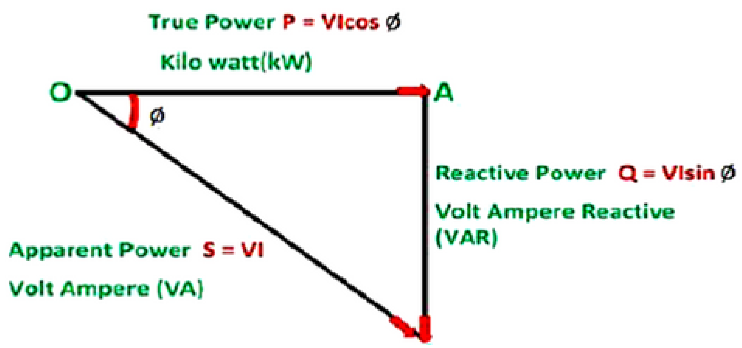
What Is Power Factor?

- It is defined as the cosine of the angle between voltage and current.
- It is dimensionless in nature.
- It is used for both single and three-phase AC circuits. It is also defined as the ratio of true or actual power to the apparent power in the ac systems.
- $\cos\phi = \frac{P}{S} = \frac{KW}{KVA}$

POWER TRIANGLE:

Power Triangle is the representation of a right angle triangle showing the relation between active power, reactive power and apparent power.

When each component of the current that is the active component (I cosϕ) or the reactive component (I sinϕ) is multiplied by the voltage V, a power triangle is obtained shown in the figure below:



From the triangle OAB

$$S = \sqrt{(P)^2 + (Q)^2}$$

$$\cos\emptyset = \frac{P}{S} = \frac{KW}{KVA}$$

Deduce Expression For Active, Reactive & Apparent Power:

ACTIVE, REACTIVE AND APPARENT POWER:

Active Power:

The power which is actually consumed or utilised in an AC Circuit is called True power or Active power or Real power. It is measured in kilowatt (kW) or MW. It is the actual outcomes of the electrical system which runs the electric circuits or load. It is denoted by "P"

Active power P = V I cos ϕ

Reactive Power:

The power which flows back and forth that means it moves in both the directions in the circuit or reacts upon itself, is called Reactive Power. The reactive power is measured in kilo volt-ampere reactive (KVAR) or MVAR. It is denoted by "Q"

Reactive power P_r or Q = V I sin ϕ

Apparent Power:

The product of root mean square (RMS) value of voltage and current is known as Apparent Power. This power is measured in KVA or MVA. It is denoted by "S"

Apparent power S = V x I = VI

REASONANCE

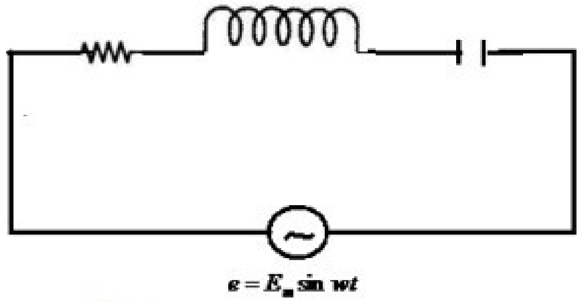
It is defined as the resonance in electrical circuit having passive or active elements represents a particular state when the current and the voltage in the circuit is maximum and minimum with respect to the magnitude of excitation at a particular frequency and the impedances being either minimum or maximum at unity power factor

Resonance are classified into two types.

(1) Series Resonance

(2) Parallel Resonance

(1) Series Resonance :- Let a resistance of 'R' ohm, inductance of 'L' henry and capacitance of 'C' farad are connected in series across A.C. supply



$e = E_m \sin wt$

The impedance of the circuit

$$Z = R + j(X_L - X_C)]$$

$$Z = \sqrt{R^2 + (X_L - X_C)^2}$$

The condition of series resonance:

The resonance will occur when the reactive part of the line current is zero

The p.f. becomes unity.

The net reactance will be zero.

The current becomes maximum.

At resonance net reactance is zero

$$X_L - X_C = 0$$

$$\Rightarrow X_L = X_C$$

$$\Rightarrow W_o L = \frac{1}{W_o C}$$

$$\Rightarrow W_o^2 LC = 1$$

$$\Rightarrow W_o^2 = \frac{1}{LC}$$

$\Rightarrow 2\pi f_o = \frac{1}{\sqrt{LC}}$

$\Rightarrow f_o = \frac{1}{2\pi\sqrt{LC}}$

Resonant frequency $(f_o) = \frac{1}{2\pi} \cdot \frac{1}{\sqrt{LC}}$

Impedance at Resonance

$Z_0 = R$

Current at Resonance

$I_o = \frac{V}{R}$

Power factor at resonance

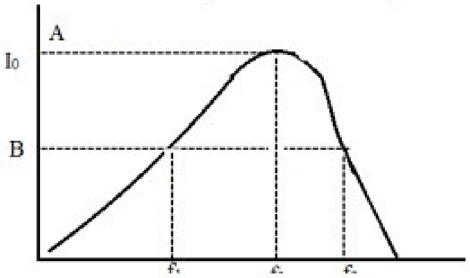
$p.f. = \frac{R}{Z_o} = \frac{R}{R} = 1 \qquad [\because Z_o = R]$

Band Width :→

At point 'A' the power loss is $I_0^2 R$.
The frequency is f_0 which is at resonance.

At point 'B' the power loss is $\frac{I_0^2 R}{2}$.

The power loss is 50% of the power loss at point 'A"/



Hence the frequencies corresponding to point 'B' is known as half power frequencies f_1 & f_2.

f_1 = Lower half power frequency

$f_1 = f_0 - \frac{R}{4\pi L}$

F_2 = Upper half power frequency

$f_2 = f_0 + \frac{R}{4\pi L}$

Band width (B.W.) is defined as the difference between upper half power frequency ad lower half power frequency.

B.W. = $f_2 - f_1 = \frac{R}{2\pi L}$

Selectivity : →

Selectivity is defined as the ratio of Band width to resonant frequency

Selectivity = $\frac{B.W.}{f_0} = \frac{R}{2\pi L}$ $\qquad Selectivity = \frac{R}{2\pi f_o L}$

Quality Factor (Q-factor) :→

It is defined as the ratio of $2\pi \times$ Maximum energy stored to energy dissipated per cycle

$$\text{Q-factor} = \frac{2\pi \times \frac{1}{2} L I_0^2}{I^2 RT}$$

$$= \frac{\pi L\left(\sqrt{2} I\right)^2}{I^2 RT} \qquad Q = \frac{Resonant\ frequency}{Bandwidth}$$

$$= \frac{\pi L.2I^2}{I^2 RT}$$

$$= \frac{\pi L.2I^2}{I^2 RT}$$

$$= \frac{2\pi L.}{RT}$$

$$\boxed{\text{Quality factor} = = \frac{2\pi f_0 L.}{R}} \qquad \left[\because = \frac{1.}{T} = f_0\right]$$

Quality factor is defined as the reciprocal of power factor.

$$\boxed{\text{Q factor} = = \frac{1.}{\cos\phi}}$$

It is the reciprocal of selectivity.

$$\text{Q-factor Or Magnification factor} = \frac{\text{Voltage across Inductor.}}{\text{Voltage across resistor}}$$

$$= \frac{I_0 X_L}{I_0 R}$$

$$= \frac{X_L}{R}$$

$$= \frac{2\pi f_0 L}{R} = \frac{W_0 L}{R}$$

$$\boxed{\text{Q- factor} = = \frac{W_0 L}{R}}$$

$$\text{Q-factor factor} = \frac{\text{Voltage across Capacotor.}}{\text{Voltage across resistor}}$$

$$= \frac{I_0 X_c}{I_0 R}$$

$$= \frac{X_C}{R}$$

$$= \frac{1}{2\pi f_0 C} = \frac{1}{2\pi f_0 CR}$$

$$\boxed{\textbf{Q-factor} = \frac{1}{W_0 CR}}$$

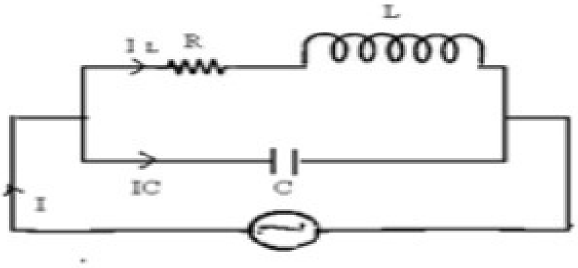
$$Q^2 = \frac{W_0 L}{R} \times \frac{1}{W_0 CR}$$

$$Q^2 = \frac{1}{R^2 C}$$

$$Q = \sqrt{\frac{1}{R^2 C}}$$

$$\boxed{Q = \frac{1}{R}\sqrt{\frac{L}{C}}}$$

(2) Parallel Resonance :- Resonance will occur when the reactive part of the line current is zero.



At resonance,

$I_C - I_L \sin\phi = 0$

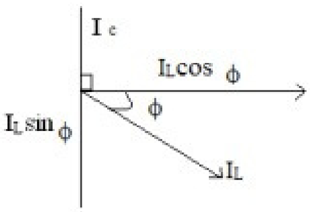
$$I_C = I_L \sin\phi$$

$$\Rightarrow \frac{V}{X_C} = \frac{V}{\sqrt{R^2 + X_L^2}} \sin\phi$$

$$\Rightarrow \frac{V}{X_C} = \frac{V}{\sqrt{R^2 + X_L^2}} \times \frac{X_L}{\sqrt{R^2 + X_L^2}}$$

$$\Rightarrow \frac{1}{X_C} = \frac{X_L}{R^2 + X_L^2}$$

$$\Rightarrow R^2 + X_L^2 = X_L . X_C$$



$$\Rightarrow Z^2 == X_L . X_C = W_0 L \times \frac{1}{W_0 C}$$

$$Z^2 = \frac{L}{C}$$

$$\Rightarrow R^2 + X_L{}^2 = \frac{L}{C}$$

$$\Rightarrow R^2 + (2\pi f_0 L)^2 = \frac{L}{C}$$

$$\Rightarrow R^2 + 4\pi^2 f_0{}^2 L^2 = \frac{L}{C}$$

$$\Rightarrow 4\pi^2 f_0{}^2 L^2 = \frac{L}{C} - R^2$$

$$\Rightarrow f_0{}^2 = \frac{1}{4\pi^2 f_0{}^2 L^2} = \left(\frac{L}{C} - R^2\right)$$

$$\Rightarrow f_0 = \frac{1}{2\pi}\sqrt{\frac{1}{LC} - \frac{R^2}{L^2}}$$

f_0 = Resonant frequency in parallel circuit.

Current at Resonance = $I_L \cos\phi$

$$= \frac{V}{\sqrt{R^2 + X_L{}^2}} . \frac{R}{\sqrt{R^2 + X_L{}^2}}$$

$$= \frac{VR}{R^2 + X_L{}^2}$$

$$= \frac{VR}{Z^2}$$

$$= \frac{VR}{L/C} = \frac{V}{L/RC}$$

$$= \frac{\text{V}}{\text{Dynamic} \quad \text{Impedence}}$$

$L/RC \rightarrow$ Dynamic Impedance of the circuit.

or, dynamic impedances is defined as the impedance at resonance frequency in parallel circuit.

Comparison of Series and Parallel Resonant Circuit :→

Item	Series ckt (R-L-C)	Parallel ckt (R– L and C)
❖ Impedance at Resonance	Minimum	Maximum
❖ Current at Resonance	Maximum= $\frac{V}{R}$	Minimum= $\frac{V}{(L/CR)}$
❖ Effective Impedance	R	$\frac{L}{CR}$
❖ P.f. at Resonance	Unity	Unity
❖ Resonant Frequency	$\frac{1}{2\pi\sqrt{LC}}$	$\frac{1}{2\pi}\sqrt{\frac{1}{LC}-\frac{R^2}{L^2}}$
❖ It Magnifies	Voltage	Current
❖ Magnification factor	$\frac{WL}{R}$	$\frac{WL}{R}$

Ex – 1: A coil of resistance 100 Ω and inductance 100 μH is connected in series with a 100 pF capacitor. The circuit is connected to a 10 V variable frequency source. Calculate (i) the resonant frequency (ii) current at resonance (iii) voltage across L and C at resonance and (iv) Q-factor of the circuit

Sol:

(i) Resonant frequency, $f_r = \frac{1}{2\pi\sqrt{LC}} = \frac{1}{2\pi\sqrt{100\times10^{-6}\times100\times10^{-12}}} = 1.59\times10^6 Hz$

(ii) Current at resonance, $I_r = \frac{V}{R} = 10/100 = 0.1A$

(iii) At resonance, $X_L = 2\pi f_r L = 2\pi\times1.59\times10^6\times100\times10^{-6} = 1000\Omega$

At resonance, $V_L = I_r X_L = 0.1\times1000 = 100V$

At resonance, $V_C = I_r X_C = 0.1\times1000 = 100V$

(iv) $Q-factor = \frac{1}{R}\sqrt{\frac{L}{C}} = \frac{1}{100}\sqrt{\frac{100\times10^{-6}}{100\times10^{-12}}} = 10$

Ex – 2: A series RLC circuit has R = 5 Ω, L = 0.2 H and C = 50 μF. The applied voltage is 200 V. Find (i) resonant frequency (ii) Q-factor (iii) bandwidth (iv) upper and lower half-power frequencies (v) current at resonance (vi) current at half-power points (vii) voltage across inductance at resonance.

Sol: $R = 5\Omega\,;\, L = 0.2H\,;\, C = 50\times10^{-6}F\,;\, V = 200volts$

(i) Resonant frequency, $f_r = \dfrac{1}{2\pi\sqrt{LC}} = \dfrac{1}{2\pi\sqrt{0.2\times 50\times10^{-6}}} = 50.33Hz$

(ii) Quality factor, $Q-factor = \dfrac{\omega_r L}{R} = \dfrac{2\pi f_r L}{R} = \dfrac{2\pi\times 50.33\times 0.2}{5} = 12.65$

(iii) Bandwidth, $BW = \dfrac{f_r}{Q-factor} = \dfrac{50.33}{12.65} = 3.98Hz$

(iv) Upper half-power frequency, $f_2 = f_r + \dfrac{BW}{2} = 50.33 + \dfrac{3.98}{2} = 52.32Hz$

Lower half-power frequency, $f_1 = f_r - \dfrac{BW}{2} = 50.33 - \dfrac{3.98}{2} = 48.34Hz$

(v) Current at resonance, $I_r = \dfrac{V}{R} = \dfrac{200}{5} = 40A$

(vi) Current at half-power points $= 0.707\times I_r = 0.707\times 40 = 28.28A$

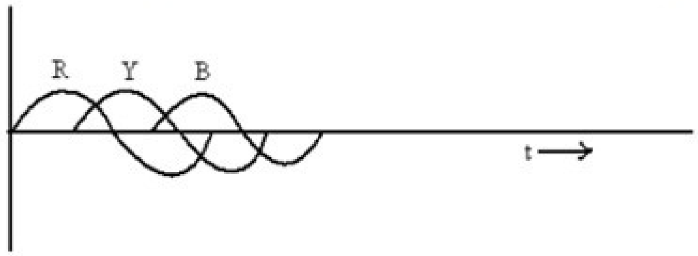
(vii) Voltage across L at resonance $I_r X_L = 40\times 2\pi\times 50.33\times 0.2 = 2529.87V$

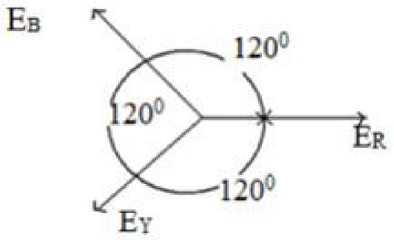
CHAPTER -06 || POLYPHASE CIRCUIT

Introduction

- **A three-phase electric system may be considered as three separate single-phase systems phase displaced from each other by 120 degree**
- **The peaks of these three phases do not occur simultaneously.**
- **Figure 1 represents the graphical pattern of voltages in a three-phase system.**
- **Polyphase systems are particularly useful for transmitting power to electric motors which rely on alternating current to rotate.**
- **The most common example is the three-phase power system used for industrial applications and for power transmission. Compared to a single-phase, two-wire system, a three phase three-wire system transmits three times as much power for the same conductor size and voltage**

Three-phase circuits consists of three windings i.e. R.Y.B



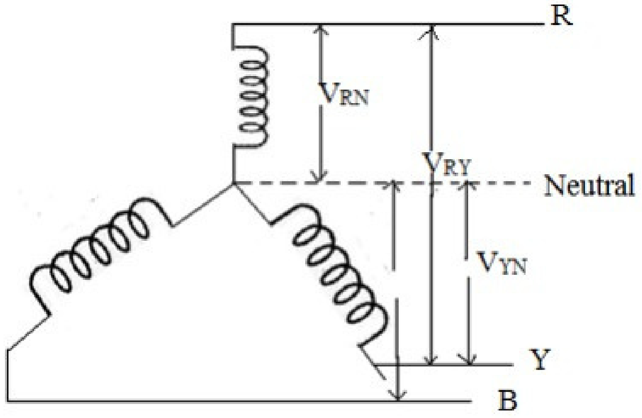


$$E_R = E_m \sin wt = E_m \angle 0$$
$$E_Y = E_m \sin(wt - 120) = E_m \angle -120$$
$$E_B = E_m \sin(wt - 240) = E_m \angle -240 = E_m \angle 120$$

3 - ϕ Circuit are divided into two types

- Star Connection
- Delta Connection

Star Connection :→



If three similar ends connected at one point, then it is known as star connected system.

The common point is known as neutral point and the wire taken from the neutral point is known as Neutral wire.

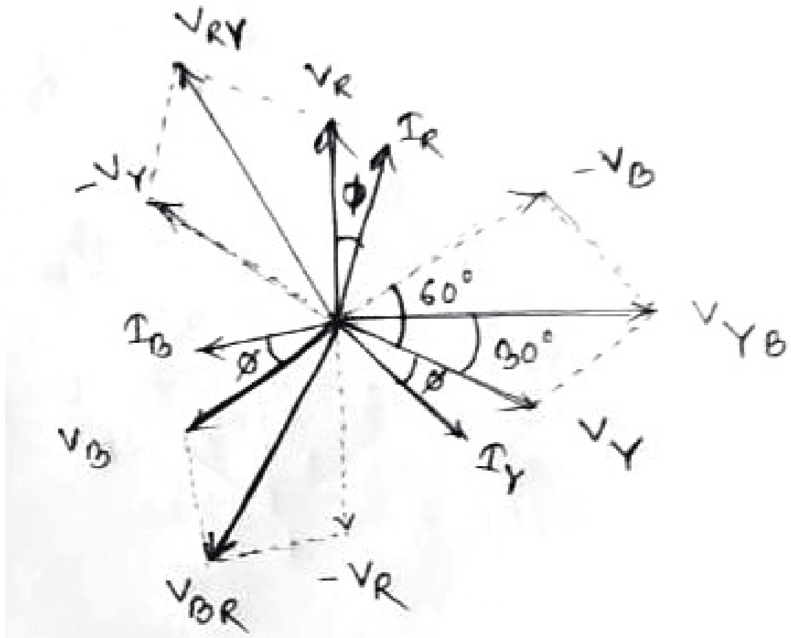
Phase Voltage :→

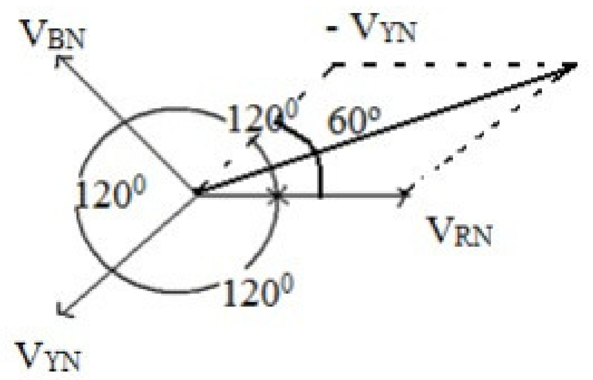
It is the potential difference between phase and Neutral.

Line Voltage : →

It is It is the potential difference between two phases.

Relation Between Phase Voltage and Line Voltage :→





Line Volatage $\vec{V_{RY}} = \vec{V_{RN}} - \vec{V_{YN}}$

$$V_L = \sqrt{V_{RN} + V_{YN} - 2V_{RN}V_{YN}Cos60^o}$$

$$= \sqrt{V_{Ph}^2 + V^2{}_{ph} - 2V_{ph}V_{ph} \times \frac{1}{2}}$$

$$= \sqrt{3V_{Ph}^2} \quad = \sqrt{3}V_{Ph}$$

$$V_L = \sqrt{3}V_{Ph}$$

Since in a balanced B –phase circuit $V_{RN} = V_{YN} = V_{BN} = V_{ph}$

Relation Between Line current and Phase Current :-

In case of star connection system the leads are connected in series with each phase

Hence the line current is equal to phase current

$I_L = I_{ph}$

Power in 3- Phase circuit:-

$P = V_{ph} I_{ph} \cos\phi$ per phase

$= 3V_{ph} I_{ph} \cos\phi$ for 3 phase

$= 3 \frac{V_L}{\sqrt{3}} I_L \cos\phi (\because V_L = \sqrt{3}V_{ph}$

$P = \sqrt{3} V_L I_L \cos\phi$

Summaries in star connection:

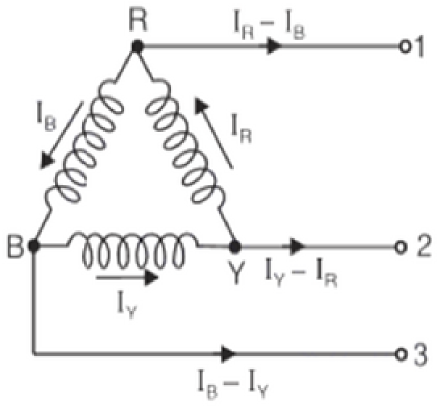
i) The line voltages are 120^0 apart from each other.

ii) Line voltages are 30^0 ahead of their respective phase voltage.

iii) The angle between line currents and the corresponding line voltage is 30+**φ**

iv) The current in line and phase are same.

Relation between Line voltage & current with Phase voltage & current in Delta connection:



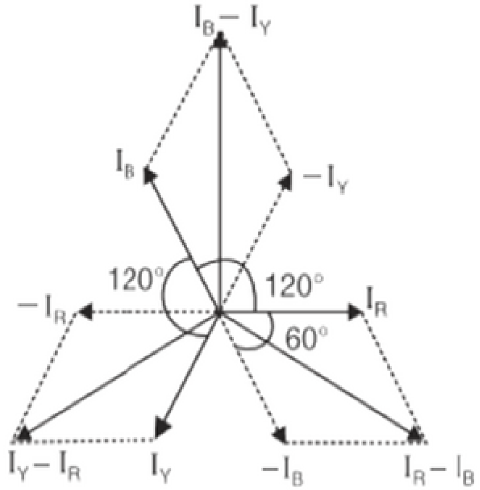


Fig. above shows a balanced 3-phase Δ - connected supply system. It is desired to find the relation between (i) line voltage and phase voltage (ii) line current and phase current.

- As one phase is included between any two lines, so magnitude of voltage between any two lines (i.e. line voltage) is equal to the magnitude of phase voltage. i.e.

  Line voltage (V_L) = Phase voltage (V_{ph})

- Since the system is balanced, the three phase currents I_R, I_Y and I_B are equal in magnitude (say I_{ph}) but displaced 120° from one another as shown in phasor diagram above.
- The current in any line is the phasor difference of the currents in the two phases attached to that line.
- Thus Current in line – 1, $= I_R - I_B$, Current in line – 2, $= I_Y - I_R$, Current in line – 3, $= I_B - I_Y$.
- The current in line – 1 (I_1) = Phasor difference of I_R and I_B

- $\therefore$
$$\begin{aligned} I_1 = I_L &= 2I_{ph}\cos\left(\theta/2\right) \\ &= 2I_{ph}\cos\left(60^\circ/2\right) \\ &= 2I_{ph}\cos 30^\circ \\ &= \sqrt{3}I_{ph} \end{aligned}$$

- So the line current, $I_L = \sqrt{3}I_{ph}$

- Line currents are 30° behind the respective phase currents.

Power equation in 3 – phase balanced circuit:

In star connection:

The total power in a star circuit is the sum of powers in the three phases. For a balanced load, the power in each phase is same.

So that, Total power, $P = 3 \times$ power in each phase $= 3 \times V_{ph} \times I_{ph} \times \cos\phi$ (if phase value is taken)

Again in star connection, $V_L = \sqrt{3}V_{ph}$ & $I_L = I_{ph}$. So that, Total power, $P = 3 \times \frac{V_L}{\sqrt{3}} \times I_L \times \cos\phi$

Or, $P = \sqrt{3}V_L I_L \cos\phi$ (KW)

Also, Reactive power is given as $Q = \sqrt{3}V_L I_L \sin\phi$ (KVAR)

$\therefore$ Total power or apparent power $(S) = \sqrt{P^2 + Q^2} = \sqrt{3}V_L I_L$

In delta connection:

The total power in a star circuit is the sum of powers in the three phases. For a balanced load, the power in each phase is same.

So that, Total power, $P = 3 \times$ power in each phase $= 3 \times V_{ph} \times I_{ph} \times \cos\phi$ (if phase value is taken)

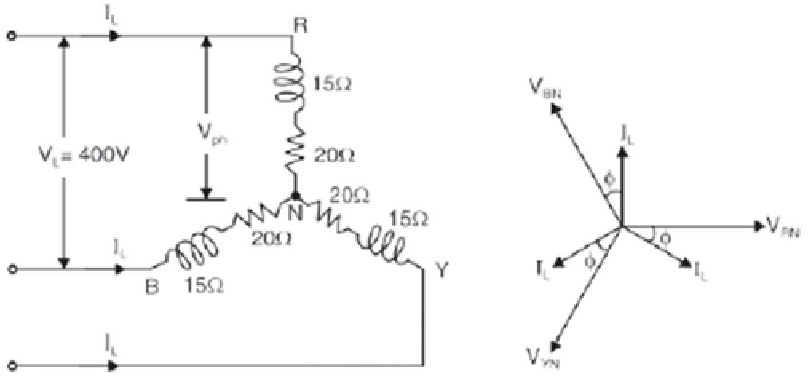
Again in star connection, $V_L = V_{ph}$ & $I_L = \sqrt{3}I_{ph}$. So that, Total power, $P = 3 \times V_L \times \frac{I_L}{\sqrt{3}} \times \cos\phi$

Or, $P = \sqrt{3}V_L I_L \cos\phi$ (KW)

Also, Reactive power is given as $Q = \sqrt{3}V_L I_L \sin\phi$ (KVAR)

$\therefore$ Total power or apparent power $(S) = \sqrt{P^2 + Q^2} = \sqrt{3}V_L I_L$

Ex – Three coils, each having a resistance of 20 Ω and an inductive reactance of 15 Ω, are connected in star to a 400 V, 3-phase, 50 Hz supply. Calculate (i) the line current (ii) power factor and (iii) power supplied.



Data given: $V_L = 400V, R_{ph} = 20\Omega, X_{Lph} = 15\Omega,\ f = 50Hz$

Sol: Phase voltage $(V_{ph}) = \frac{V_L}{\sqrt{3}} = \frac{400}{\sqrt{3}} = 231\text{V}$

Impedance/phase, $Z_{ph} = \sqrt{R_{ph}^2 + X_{Lph}^2} = \sqrt{20^2 + 15^2} = 25\Omega$

Phase current, $I_{ph} = \frac{V_{ph}}{I_{ph}} = \frac{231}{25} = 9.24A$

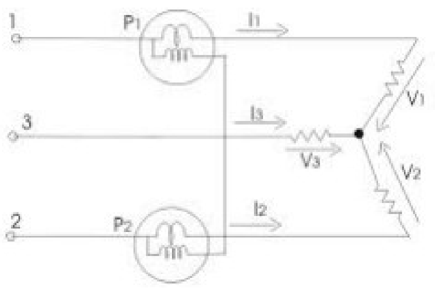
∴ **Line current,** $I_L = I_{ph} = 9.24A$

∴ **Power factor,** $\frac{R_{ph}}{Z_{ph}} = \frac{20}{25} = 0.8lag$

∴ **Power supplied,** $P = \sqrt{3}V_L I_L \cos\phi = \sqrt{3} \times 400 \times 9.24 \times 0.8 = 5121W$

Or, $P = 3I_{ph}^2 R_{ph} = 3 \times (9.24)^2 \times 20 = 5121W$

Measurement of power By Two Watt Meter Method :-



Phasor Diagram :-

Let V_R, V_Y,V_B are the r.m.s value of 3-ϕ voltages and I_R,I_Y,I_B are the r.m.s. values of the currents respectively.

Current in R-phase which flows through the current coil of watt-meter

$W_1 = I_R$

And $W_2 = I_Y$

Potential difference across the voltage coil of $W_1 = \vec{V}_{RB} = \vec{V}_R - \vec{V}_B$

And $W_2 = \vec{V}_{YB} = \vec{V}_Y - \vec{V}_B$

Assuming the load is inductive type watt-meter W_1 reads.

$$W_1 = V_{RB} I_R \cos(30-\phi)$$

$$W_1 = V_L I_L \cos(30-\phi) \text{---------------------------}(1)$$

Wattmeter W_2 reads

$$W_2 = V_{YB} I_Y \cos(30+\phi)$$

$$W_2 = V_L I_L \cos(30+\phi) \text{---------------------------}(2)$$

$$W_1 + W_2 = V_L I_L \cos(30-\phi) + V_L I_L \cos(30+\phi)$$

$$= V_L I_L [\cos(30-\phi) + V_L I_L \cos(30+\phi)]$$

$$= V_L I_L (2\cos 30^o \cos\phi)$$

$$= V_L I_L (2 \times \frac{\sqrt{3}}{2} \cos\phi)$$

$$W_1 + W_2 = \sqrt{3} V_L I_L \cos\phi \text{------------------------------}(3)$$

$$W_1 - W_2 = V_L I_L [\cos(30-\phi) - \cos(30+\phi)$$

$$= V_L I_L (2\sin 30^o \sin\phi)$$

$$= V_L I_L (2\times\frac{1}{2}\times\sin\phi)$$

$$W_1 - W_2 = V_L I_L \sin\phi$$

$$\frac{W_1 - W_2}{W_1 + W_2} = \frac{V_L I_L \sin\phi}{\sqrt{3} V_L I_L \cos\phi}$$

$$\frac{1}{\sqrt{3}} = \tan\phi$$

$$\Rightarrow \tan\phi = \sqrt{3}\left(\frac{W_1 - W_2}{W_1 + W_2}\right)$$

$$\Rightarrow \phi = \tan^{-1}\sqrt{3}\left(\frac{W_1 - W_2}{W_1 + W_2}\right)$$

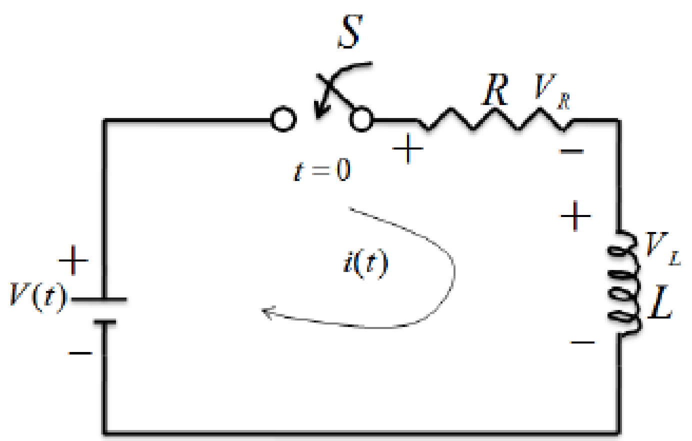
Variation in wattmeter reading with respect to p.f:

Pf	W_1 reading	W_2 reading
φ=0,cos **φ**=1	+ve equal	+ve equal
φ=60,cos **φ**=0.5	0	+ve
φ=90,cos **φ**=0	-ve, equal	+ve equal

CHAPTER -07 || TRANSIENTS

Whenever a network containing energy storage elements such as inductor or capacitor is switched from one condition to another, either by change in applied source or change in network elements, the response current and voltage change from one state to the other state. The time taken to change from an initial steady state to the final steady state is known as the transient period. This response is known as transient response or transients. The response of the network after it attains a final steady value is independent of time and is called the steady-state response. The complete response of the network is determined with the help of a differential equation.

Response to R-L circuit under DC condition:



We know that, $$i=\frac{v}{R}, V_L=L\frac{di}{dt}, i_C=C\frac{dv}{dt}$$

In the circuit above applying KVL, we have,

$$V_R+V_L=V$$

$$=i(t)R+L\frac{di(t)}{dt}=v(t)$$

$$=Ri(t)+L\frac{di(t)}{dt}=v(t)$$

Now divide L in both the sides, we get,

$$\frac{di(t)}{dt} + \frac{R}{L} i(t) = \frac{v(t)}{L} \quad \text{.............................. (1).}$$

The eq. (1) is a linear differential equation in 1st order. $\frac{di}{dt} + Pi = Q$

The solution of the linear differential equation is given as,

$$C.F. + P.I.$$

$C.F =$

$$\frac{di(t)}{dt} + \frac{R}{L} i(t) = 0$$

Or, $\frac{di(t)}{dt} = -\frac{R}{L} i(t)$

Now integrating both the side, we get,

Or, $\int \frac{di}{i} = -\frac{R}{L} \int dt$

$\log i = -\frac{R}{L} t + \mathrm{k}$

Or, $e^{\log i} = Ce^{-\frac{R}{L}t}$

$i(t) = Ce^{-\frac{R}{L}t}$

$P.I =$

Put $t = \infty$

$i(t) = \frac{V}{R}$

The solution of the linear differential equation is given as,

$$C.F. + P.I.$$

$i(t) = C.F + P.I$

$$= Ce^{-\frac{R}{L}t} + \frac{V}{R}$$

Put $t = 0$, as it is the starting point

$$0 = Ce^{-0} + \frac{V}{R}$$

$$C = -\frac{V}{R}$$

Now the current $i(t)$ is given as,

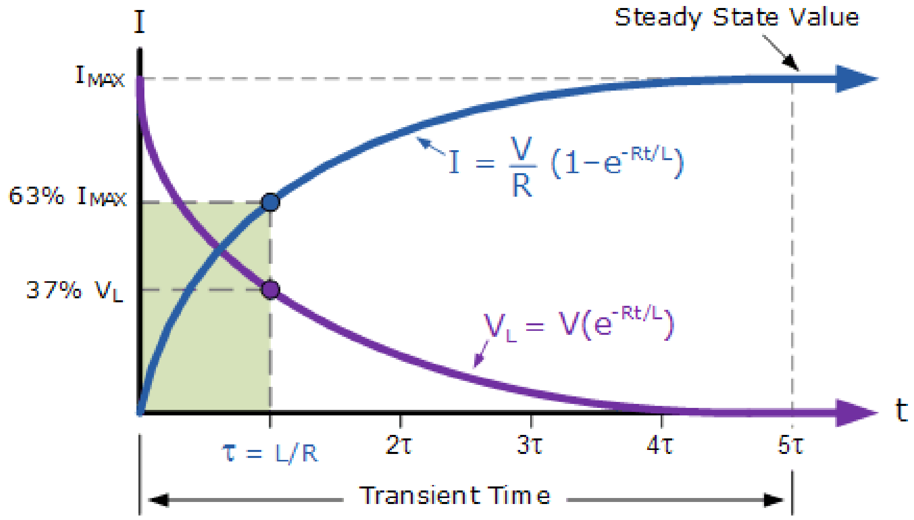
$$i(t) = \frac{V}{R} - \frac{V}{R}e^{-\frac{t}{\lambda}}$$

$$= \frac{V}{R}\left(1 - e^{-\frac{t}{\lambda}}\right)$$

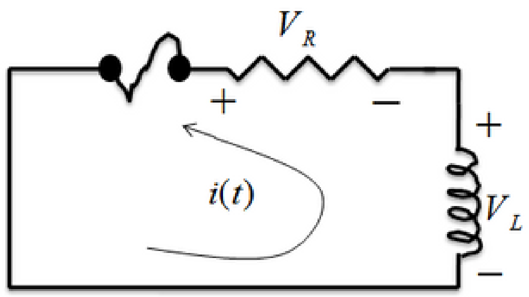
$\therefore$ $$i(t) = I_0\left(1 - e^{-\frac{t}{\lambda}}\right)$$ → **charging equation of current.**

Where λ = time constant and is given by,

$$\lambda = \frac{L_{eq}}{R_{eq}}$$



Discharging response in R-L circuit



Now the inductor starts discharging through the resistance and by applying **KVL** we have,

$$V_R + V_L = 0$$

Or, $$i(t)R + L\frac{di(t)}{dt} = 0$$

Now divide ***L*** in both the sides, we get,

$$\frac{di(t)}{dt} + \frac{R}{L}i(t) = 0 \quad \text{.................................. (1).}$$

The eq. (1) is a linear differential equation in 1st order. $\frac{di}{dt} + Pi = Q$

The solution of the linear differential equation is given as,

$$C.F. + P.I.$$

$$C.F =$$

$$\frac{di(t)}{dt} + \frac{R}{L}i(t) = 0$$

Or, $$\frac{di(t)}{dt} = -\frac{R}{L}i(t)$$

Now integrating both the side, we get,

Or, $$\int \frac{di}{i} = -\frac{R}{L}\int i(t)$$

$$\log i = \frac{R}{L}t + \mathrm{k}$$

Or, $e^{\log i} = Ce^{-\frac{R}{L}t}$

$$i(t) = Ce^{-\frac{R}{L}t}$$

$P.I =$

If $Q = 0$, $P.I = 0$

The solution of the linear differential equation is given as,

$$C.F. + P.I.$$

$$i(t) = C.F + P.I$$

$$= Ce^{-\frac{R}{L}t}$$

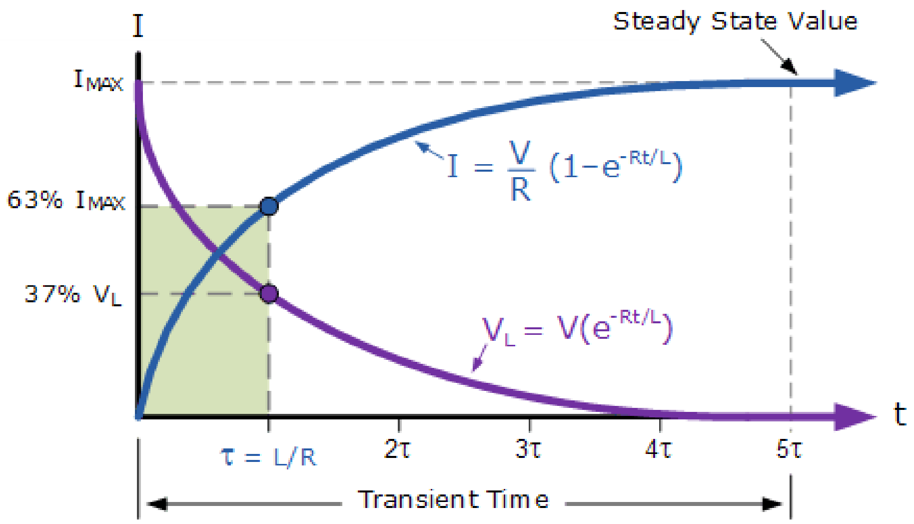
Applying initial condition at $t = 0$,

$$\frac{V}{R} = Ce^{-0}$$

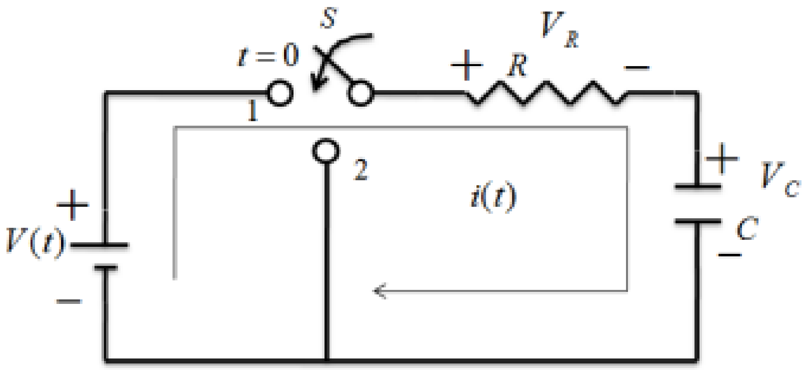
$$C = \frac{V}{R}$$

Now the current $i(t)$ is given as,

$$i(t) = \frac{V}{R}e^{-\frac{t}{\lambda}}$$



Charging response in R-C circuit



When the switch (S) in position 1 at $t = 0$,

By applying ***KVL,*** We get,

$$V = V_R + V_C$$

Or, $$IR + V_C = V$$

Or, $$C\frac{dV_C}{dt}R + V_C = V$$

$$\because\ i_C = C\frac{dv}{dt}$$

Or, $$dv = \frac{1}{C}\int i_C dt$$

Or, $$V_C = \frac{1}{C}\int i_C dt$$

Or, $$\frac{dV_C}{dt} + \frac{V_C}{RC} = \frac{V}{RC} \text{ .... (1)}$$ ($\because$ Dividing R on both the sides).

The above equation (1) is a first order differential equation and in the form of $\frac{dv}{dt} + pv = Q$.

Now the solution of the above eq. (1) is given by,

$$C.F. + P.I.$$

$C.F =$

$$\frac{dV_C}{dt} + \frac{V_C}{RC} = 0$$

$$\frac{dV_C}{dt} = -\frac{V_C}{RC}$$

Or,

$$\frac{dV_C}{V_C} = -\frac{1}{RC}dt$$

Now integrating both the sides, we have,

$$\int \frac{dV_C}{V_C} = -\frac{1}{RC}\int dt$$

$$\log V_C = -\frac{t}{RC} + \mathrm{k}$$

$$V_C = Ce^{-\frac{t}{RC}}$$

$P.I =$

At $t = \infty, V_C = V$

The net solution is

$$C.F. + P.I.$$

$$V_C = V + Ce^{-\frac{t}{RC}}$$

Now, putting the initial condition at $t = 0, V_C = 0$

Now, putting the initial condition at $t = 0, V_C = 0$

$$0 = V + Ce^{-0}$$

Or, $$0 = V + C$$

Or, $$C = -V$$

Now putting the value of C in above equation, we get,

$$V_C(t) = V + (-V)e^{-\frac{t}{RC}}$$

Or, $$V_C(t) = V - Ve^{-\frac{t}{RC}}$$

Again, $$\lambda = RC$$

So, $$V_C(t) = V_0\left(1 - e^{-\frac{t}{\lambda}}\right)$$

Where $\lambda =$ time constant and is given by,

$\lambda = R_{eq} \times C_{eq}$

Now net current or capacitor current is given as,

$$i_C(t) = C\frac{dV_C}{dt}$$

$$= C\frac{d}{dt}\left[V_0\left(1-e^{-\frac{t}{\lambda}}\right)\right]$$

$$= CV_0\frac{d}{dt}\left[\left(1-e^{-\frac{t}{\lambda}}\right)\right]$$

$$= CV_0\left[0-e^{-\frac{t}{RC}}\times\left(-\frac{1}{RC}\right)\right]$$

$$= \frac{V_0}{R}e^{-\frac{t}{RC}}$$

$\therefore$ $$i(t) = I_0 e^{-\frac{t}{\lambda}}$$

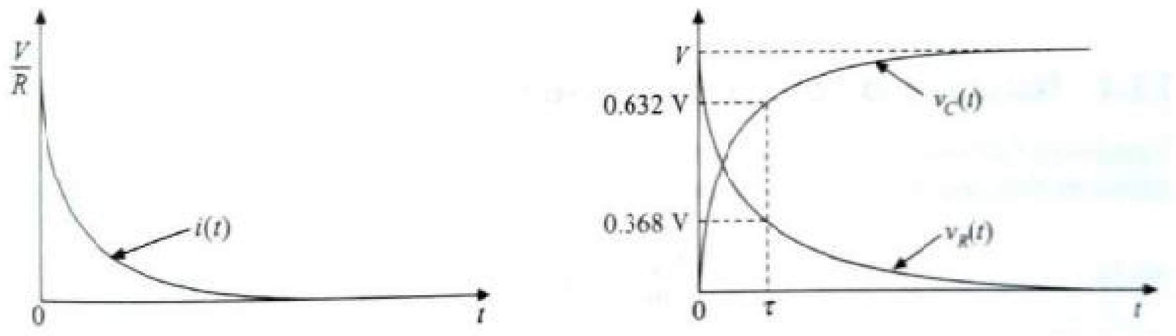
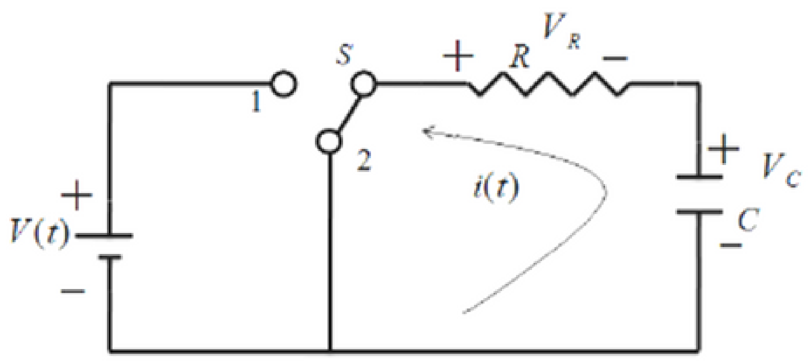


Fig : Transient current and voltages in RC circuit with DC excitation.

❖ Discharging response in R-C circuit



When the switch moves from (1) to (2), the capacitor starts discharging.

Now by applying ***KVL***, We get,

$$V_R + V_C = 0$$

Or, $$i(t)R + V_C = 0$$

Or, $$C\frac{dV_C(t)}{dt} \times R + V_C = 0 \qquad \because \; i(t) = C\frac{dV_C(t)}{dt}$$

Now dividing ***RC*** on both the sides, we have,

Or, $$\frac{dV_C(t)}{dt} + \frac{1}{RC}V_C = 0 \text{ .... (1).}$$

The above equation (1) is in the form of first order differential equation i.e. $\frac{dv}{dt} + pv = Q$

∴ Now the solution of the above eq. (1) is given by,

$$C.F. + P.I.$$

$C.F =$

$$\frac{dV_C(t)}{dt} + \frac{V_C}{RC} = 0$$

Or, $$\frac{dV_C(t)}{dt} = -\frac{1}{RC}V_C(t)$$

Now integrating both the sides,

Or, $$\int \frac{dV_C(t)}{V_C(t)} = -\frac{1}{RC}\int dt$$

Or, $$\log V_C(t) = -\frac{t}{RC} + \mathrm{k}$$

Or, $$V_C(t) = Ce^{-\frac{t}{RC}}$$

$$P.I =$$

At $Q = 0, P.I = 0$

Now the solution is given as,

$$C.F. + P.I.$$

$$V_C(t) = Ce^{-\frac{t}{RC}} + 0$$

Now putting the initial condition at $t = 0$

$$V = Ce^{-0}$$

Or, $$C = V$$

$\therefore$ $$V_C(t) = V_0 e^{-\frac{t}{RC}} volt$$

Now $$i(t) = C\frac{dV_C(t)}{dt}$$

$$= C\frac{d}{dt}\left(V_0 e^{-\frac{t}{RC}}\right)$$

$$= CV_0\frac{d}{dt}\left(e^{-\frac{t}{RC}}\right)$$

$$= CV_0\left(-\frac{1}{RC}\right) \times e^{-\frac{t}{RC}}$$

$$= -\frac{V_0}{R}e^{-\frac{t}{RC}}$$

$$\therefore \qquad i(t) = -I_0 e^{-\frac{t}{\lambda}} Amp$$

Where λ = time constant and is given by,

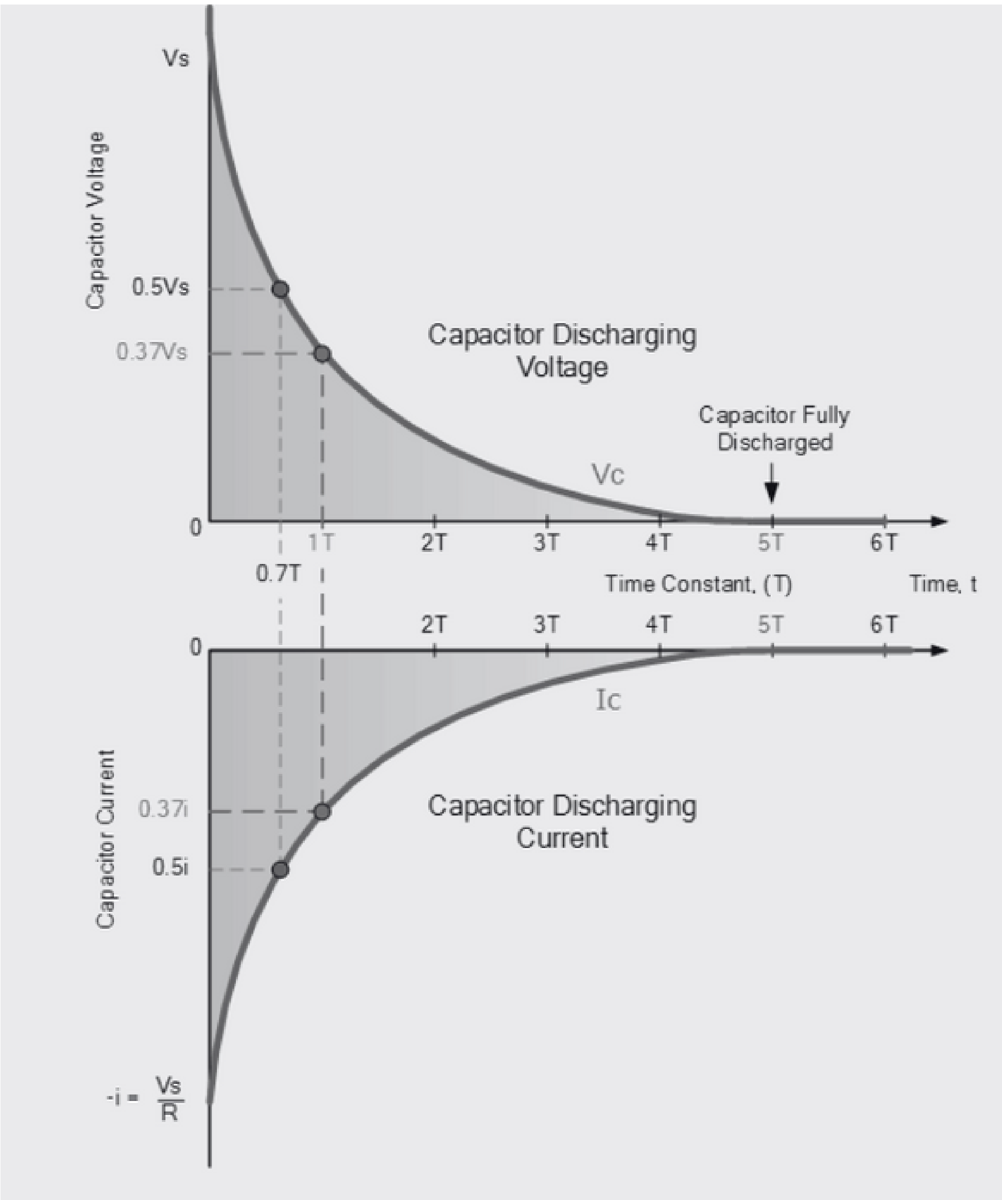
$$\lambda = R_{eq} \times C_{eq}$$

$\therefore$ Voltage across resistance $V_R(t)$ is given as,

$$V_R(t) = i(t) \times R$$

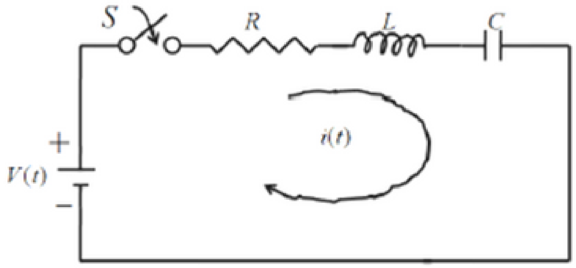
$$= -I_0 e^{-\frac{t}{\lambda}} \times R$$

$$V_R(t) = -V_0 \times e^{-\frac{t}{\lambda}} volt$$



❖ **DC response of an R-L-C circuit:**

Consider a circuit consisting of a resistance, inductance and capacitance as shown in fig. The capacitor and inductor in the circuit initially uncharged and are in series with the resistor. When the switch S is closed at $t = 0$, we can find the complete solution for the current. Application of Kirchhoff's voltage law to the circuit results the following differential equation.



Applying ***KVL*** to the above circuit, we have, $V_R + V_L + V_C = V$

$$V = Ri + L\frac{di}{dt} + \frac{1}{C}\int idt \text{ .... (1)}$$

By differentiating the above equation w.r.t. **t**, we get,

$$0 = R\frac{di}{dt} + L\frac{d^2t}{dt^2} + \frac{1}{C}i$$

Or, $$\frac{d^2t}{dt^2} + \frac{R}{L}\frac{di}{dt} + \frac{1}{LC}i = 0 \text{ .... (2)}$$ (Dividing ***L*** in all)

The above equation-(2) is a second order differential equation with only the complementary function. The particular solution for the above equation is zero. The characteristics equation for this type of differential equation is

$$\left(D^2 + \frac{R}{L}D + \frac{1}{LC}\right) \times i = 0 \text{ .... (3)} \qquad \left(\because \frac{d}{dt}\right) = D$$

Now the roots of the equations-(3) are D_1 & D_2

$$D_1, D_2 = -\frac{R}{2L} \pm \sqrt{\left(\frac{R}{2L}\right)^2 - \frac{1}{LC}}$$

Let $\alpha = \frac{R}{2L}$ & $\beta = \sqrt{\left(\frac{R}{2L}\right)^2 - \frac{1}{LC}}$

So $D_1 = -\alpha + \beta$ and $D_2 = -\alpha - \beta$

Where $\omega = \frac{1}{\sqrt{LC}} \rightarrow$ **Natural frequency,** $\alpha = \frac{R}{2L} \rightarrow$ **Damping coefficient and**

$\frac{\alpha}{\omega} = \xi = \frac{R}{2}\sqrt{\frac{C}{L}} \rightarrow$ **Damping factor**

Case-1: If $\left(\frac{R}{2L}\right)^2 \rangle \frac{1}{LC} \rightarrow$ over damped response (roots are negative, real and unequal).

So $D_1 = -\alpha + \beta$ and $D_2 = -\alpha - \beta$

The solution of the above equation is

$$i = C_1 e^{-D_1 t} + C_2 e^{-D_2 t}$$

Case-2: If $\left(\frac{R}{2L}\right)^2 = \frac{1}{LC} \rightarrow$ critically damped response (roots are negative, real and equal).

So $D_1 = D_2 = -\alpha$

The solution of the above equation is

$$i = C_1 e^{-\alpha t} + C_2 e^{-\alpha t}$$

Case-3: If $\left(\frac{R}{2L}\right)^2 \langle \frac{1}{LC} \rightarrow$ under damped response (roots will be complex conjugate).

So $D_1 = -\alpha + j\beta$ and $D_2 = -\alpha - j\beta$

The solution of the above equation is

$$i = e^{-\alpha t}\left[C_1 \cos \beta t + C_2 \sin \beta t\right]$$

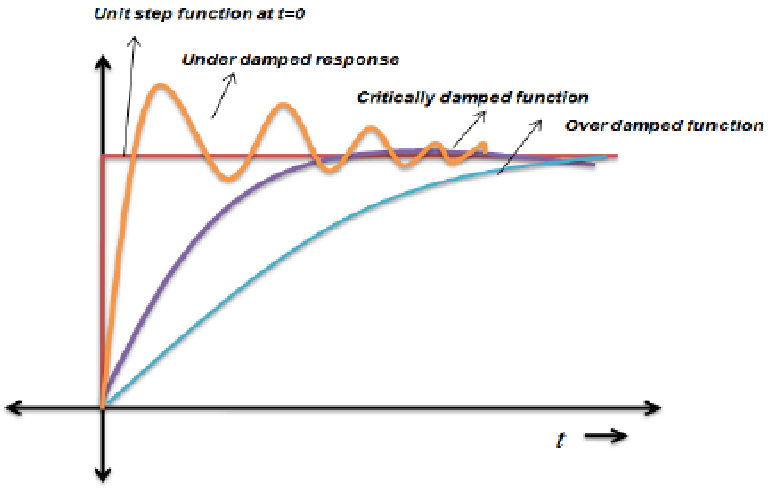
Case-4: If $R = 0 \rightarrow$ only oscillation response (roots will be imaginary).

So $D_1 = j\sqrt{\frac{1}{LC}} = j\omega$ and $D_2 = -j\omega$

The solution of the above equation is

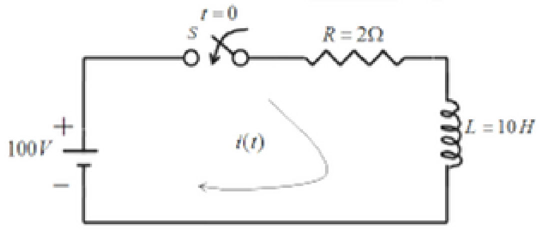
$$i = \left[C_1 \cos \omega t + C_2 \sin \omega t\right]$$

Now the responses of above four cases are shown in fig. below.



Examples:

Q.1. Find the current in a series ***R-L*** circuit having ***R=2Ω & L=10H,*** while a dc voltage of ***100V*** is applied. What is the value of this current after 5sec of switching **ON**?



Ans:

We know that, the charging current equation is given as,

$$i(t) = I_0\left(1 - e^{-\frac{t}{\lambda}}\right)$$

λ = time constant and is given as,

$$\lambda = \frac{L_{eq}}{R_{eq}} = \frac{10}{2} = 5\sec$$

Steady state current $I_0 = I_{S.S} = \frac{V_0}{R} = \frac{100}{2} = 50A$

$\therefore$ $$i(t) = I_0\left(1 - e^{-\frac{t}{\lambda}}\right)$$

$$= 50\left(1 - e^{-\frac{5}{5}}\right)$$

$$= 50\left(1 - e^{-1}\right)$$

$$= 50(1 - 0.3678)$$

$$= 31.61A$$

$$\therefore \qquad i(t) = 31.61A$$

Q.2. A dc voltage of ***100V*** is applied to a coil of ***R=10Ω & L=20H.*** What is the value of current 0.2 sec after switching ***ON*** and the time taken for the current to reach one half of its final value?

Ans:

We know that, the charging current equation is given as,

$$i(t) = I_0\left(1 - e^{-\frac{t}{\lambda}}\right)$$

λ = time constant and is given as,

$$\lambda = \frac{L_{eq}}{R_{eq}} = \frac{20}{10} = 2\sec$$

Steady state current $I_0 = I_{S.S} = \frac{V_0}{R} = \frac{100}{10} = 10A$

$$\therefore \qquad i(t) = I_0\left(1 - e^{-\frac{t}{\lambda}}\right)$$

$$= 10\left(1 - e^{-\frac{0.2}{2}}\right)$$

$$= 10\left(1 - e^{-0.1}\right)$$

$$= 10(1 - 0.9048)$$

$$= 0.952A$$

$$\therefore \qquad i(t) = 0.952A$$

Again time taken **(t)** to reach half of I_0 i.e. $\frac{I_0}{2} = \frac{10}{2} = 5A$ is given as,

$$5 = 10\left(1 - e^{-\frac{t}{0.2}}\right)$$

Or, $$1 - e^{-\frac{t}{0.2}} = \frac{5}{10}$$

Or, $$1 - 0.5 = e^{-\frac{t}{0.2}}$$

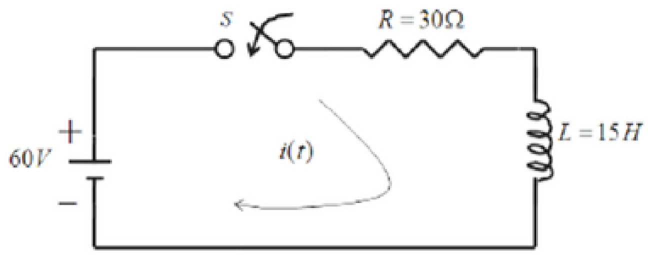
Or, $$\ln\left(e^{-\frac{t}{0.2}}\right) = \ln(0.5)$$

Or, $$-\frac{t}{0.2} = -0.6931$$

Or, $$t = 0.2 \times 0.6931 = 0.1386$$

$\therefore$ $$t = 0.1386\,\text{sec}$$

Q.3. A series ***R-L*** circuit with ***R=30Ω & L=15H*** has a constant voltage ***V=60V*** applied at ***t=0*** as shown in fig. below. Determine the current ***i,*** the voltage across resistor and across inductor.



Ans:

We know that, the charging current equation is given as,

$$i(t) = I_0\left(1 - e^{-\frac{t}{\lambda}}\right)$$

λ = time constant and is given as,

$$\lambda = \frac{L_{eq}}{R_{eq}} = \frac{15}{30} = 0.5\,\text{sec}$$

Steady state current $I_0 = I_{S.S} = \frac{V_0}{R} = \frac{60}{30} = 2A$

$$\therefore \qquad i(t) = I_0\left(1 - e^{-\frac{t}{\lambda}}\right)$$

$$= 2\left(1 - e^{-\frac{t}{0.5}}\right)$$

$$= 2\left(1 - e^{-2t}\right)A$$

$$\therefore \qquad i(t) = 2\left(1 - e^{-2t}\right)Amp.$$

$\therefore$ Voltage across the resistor $(V_R) =$

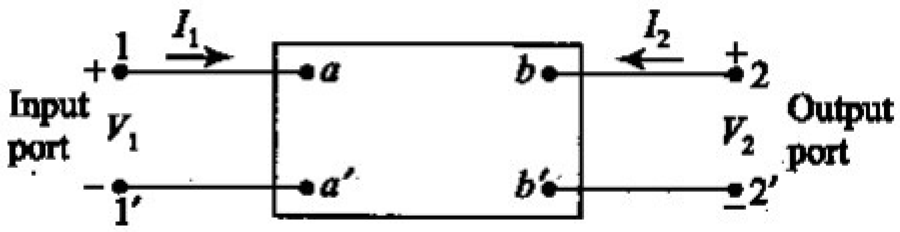
$$V_R = i(t) \times R = 2\left(1 - e^{-2t}\right) \times 30 = 60 \times \left(1 - e^{-2t}\right)volt$$

$\therefore$ Voltage across the inductor $(V_L) =$

$$V_L = V_0 \times e^{-2t} = 60 \times e^{-2t} volt$$

CHAPTER -08 || TWO-PORT NETWORK:

- Generally, any network may be represented schematically by a rectangular box. A network may be used for representing either Source or Load , or for a variety of purposes.
- A pair of terminals at which a signal may enter or leave a network is called a port.
- A port is defined as any pair of terminals into which energy is withdrawn ,or where the network variables may be measured



Types of parameters:
Z Parameter
Y Parameter
h-Parameter
ABCD Parameter

OPEN CIRCUIT IMPEDANCE (Z) PARAMETERS:

- We can also calculate the impedance parameters after making two sets of measurements.

$$V_1 = z_{11} I_1 + z_{12} I_2$$

$$V_2 = z_{21} I_1 + z_{22} I_2$$

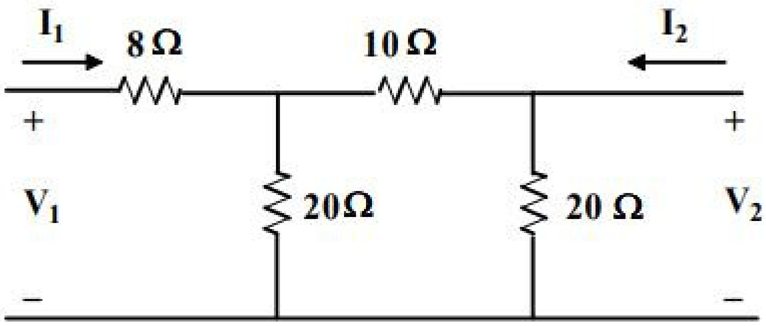
If the right port is an open circuit (I_2=0), then we can easily solve for two of the impedance parameters: Similarly by open circuiting left hand port ($I_{1=0}$) we can solve for the other two parameters.

$$Z_{11} = inputimpedence = \frac{V_1}{I_1} | I_2 = 0 \qquad Z_{21} = forwardtransferimpedence = \frac{V_2}{I_1} | I_2 = 0$$

$$Z_{12} = reversetransferimpedence = \frac{V_1}{I_2} | I_1 = 0 \quad Z_{22} = outputimpedence = \frac{V_2}{I_2} | I_1 = 0$$

Example:

Given the following circuit. Determine the Z parameters.



$$Z_{11} = 8 + 20||30 = 20\ \Omega$$

$$Z_{22} = 20||30 = 12\ \Omega$$

$$Z_{12} = \frac{V_1}{I_2}|I_1 = 0$$

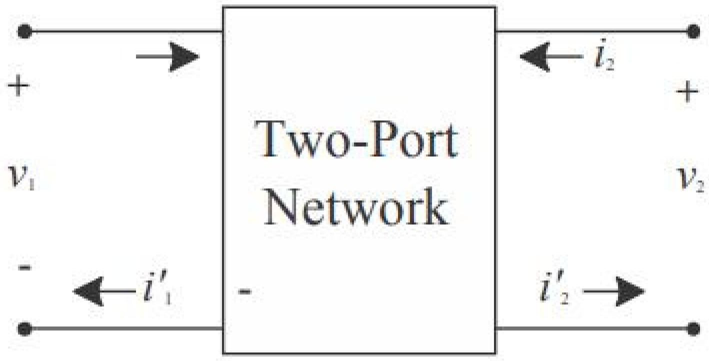
$$V_1 = \frac{20xI_2x20}{20+30} = 8xI_2 \quad \text{Therefore } z_{12} = \frac{8xI_2}{I_2} = 8\,\Omega = z_{21}$$

The Z parameter equations can be expressed in matrix form as follows.

$$\begin{bmatrix} V_1 \\ V_2 \end{bmatrix} = \begin{bmatrix} z_{11} & z_{12} \\ z_{21} & z_{22} \end{bmatrix} \begin{bmatrix} I_1 \\ I_2 \end{bmatrix}$$

$$\begin{bmatrix} V_1 \\ V_2 \end{bmatrix} = \begin{bmatrix} 20 & 8 \\ 8 & 12 \end{bmatrix} \begin{bmatrix} I_1 \\ I_2 \end{bmatrix}$$

SHORT-CIRCUIT ADMITTANCE (Y) PARAMETERS:



$$I_1 = y_{11}V_1 + y_{12}V_2$$

$$I_2 = y_{21}V_1 + y_{22}V_2$$

$$Y_{11} = \text{input admittance} = \frac{I_1}{V_1}\Big|V_2 = 0$$

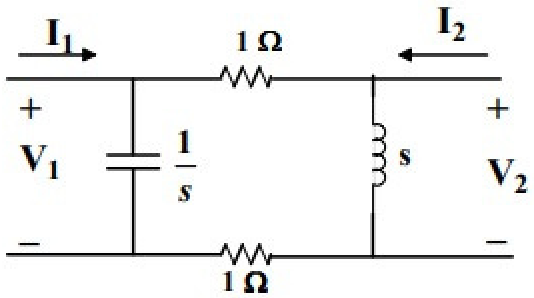
$$Y_{21} = \text{forward transfer admittance} = \frac{I_2}{V_1}\Big|V_2 = 0$$

$$Y_{22} = \text{output admittance} = \frac{I_1}{V_2}\Big|V_1 = 0$$

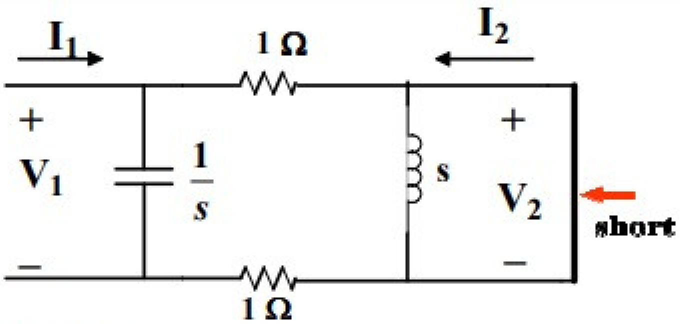
$$Y_{12} = \text{reverse transfer admittance} = \frac{I_1}{V_2}\Big|V_1 = 0$$

Example:

Given the following circuit. Determine the Y parameters.



$$I_1 = y_{11}V_1 + y_{12}V_2$$
$$I_2 = y_{21}V_1 + y_{22}V_2$$



To find y_{11}

$$V_1 = I_1\left(\frac{2/s}{2+1/s}\right) = I_1\left[\frac{2}{2s+1}\right]$$

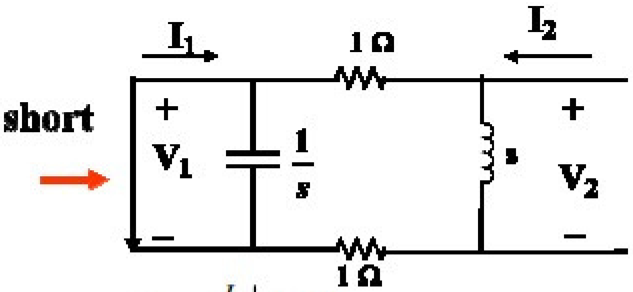
$$\text{So } y_{11} = \frac{I_1}{V_1}\; y_{11} = \frac{I_1}{V_1}\bigg|_{V_2=0} = s+0.5$$

To find y_{12} and y_{21} we reverse things and short V_1

$$y_{21} = \frac{I_2}{V_1}\Big| V_2 = 0$$

$$V_1 = -2I_2$$

$$y_{21} = \frac{I_2}{V_1} = 0.5 \text{ S}$$



$$y_{12} = \frac{I_1}{V_2}\Big| V_1 = 0$$

$$V_2 = -2I_1 \quad y_{12} = \frac{I_1}{V_2} = 0.5s$$

$$y_{22} = 0.5 + \frac{1}{s}$$

$$y_{22} = \frac{I_2}{V_2}\Big|V_1 = 0 \qquad V_2 = I_2 \frac{2s}{(s+2)} \qquad y_{22} = 0.5 + \frac{1}{s}$$

Transmission (ABCD) parameters:

$$V_1 = AV_2 - BI_2$$

$$I_1 = CV_2 - DI_2$$

$$A = \frac{V_1}{V_2},\ when\ I_2 = 0$$

$$B = -\frac{V_1}{I_2},\ when\ V_2 = 0$$

$$C = \frac{I_1}{V_2},\ when\ I_2 = 0$$

$$D = -\frac{I_1}{I_2},\ when\ V_2 = 0$$

Hybrid h-parameter

$$V_1 = h_{11}I_1 + h_{12}V_2$$

$$I_2 = h_{21}I_1 + h_{22}V_2$$

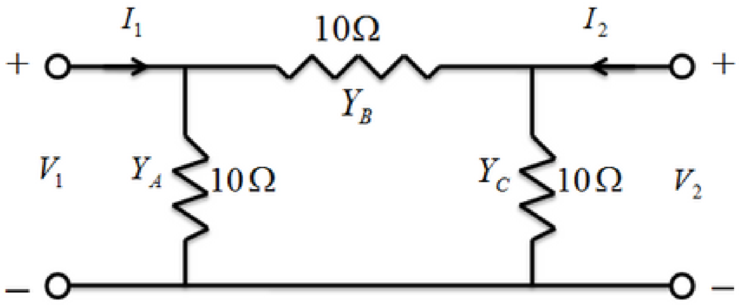
h_{11} = input impedance = $\frac{V_1}{I_1}\Big|V_2 = 0$

h_{21} = forward current ratio = $\frac{I_2}{I_1}\Big|V_2 = 0$

h_{12} = reverse voltage ratio = $\frac{V_1}{V_2}\Big|I_1 = 0$

h_{22} = output admittance = $\frac{I_2}{V_2}\Big|I_1 = 0$

Q.Find the Y-parameters of the network?



At node-A, **by KCL,**

$$I_1 = \frac{V_1}{10} + \frac{V_1 - V_2}{10} = \frac{V_1}{10} + \frac{V_1}{10} - \frac{V_1}{10}$$

$$= V_1\left(\frac{1}{10} + \frac{1}{10}\right) - \left(\frac{1}{10}\right)V_2$$

$$I_1 = 0.2V_1 - 0.1V_2 \quad \text{.............................} \mathbf{(1)}$$

At node-B, **by KCL,**

$$I_2 = \frac{V_2}{10} + \frac{V_2 - V_1}{10} = \frac{V_2}{10} + \frac{V_2}{10} - \frac{V_1}{10}$$

$$= V_2\left(\frac{1}{10} + \frac{1}{10}\right) - \left(\frac{1}{10}\right)V_1$$

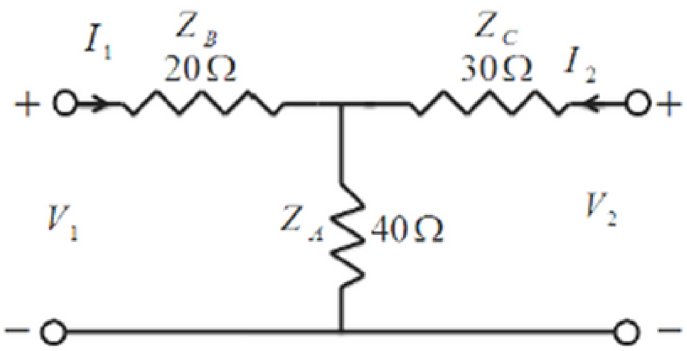
$$I_2 = -0.1V_1 + 0.2V_2 \quad \text{.............................} \mathbf{(2)}$$

Now comparing eq. **1 and 2**, we get,

$$Y_{11} = 0.2, Y_{12} = -0.1$$

$$Y_{21} = -0.1, Y_{22} = 0.2$$

Q. Determine the Z-parameters of the network?



Solution:

As it is a ***T*** network, so we can directly find the Z-parameter.

We know that,

$$V_1 = Z_{11}I_1 + Z_{12}I_2 \quad \text{.................................} \quad (1)$$

$$V_2 = Z_{21}I_1 + Z_{22}I_2 \quad \text{.................................} \quad (2)$$

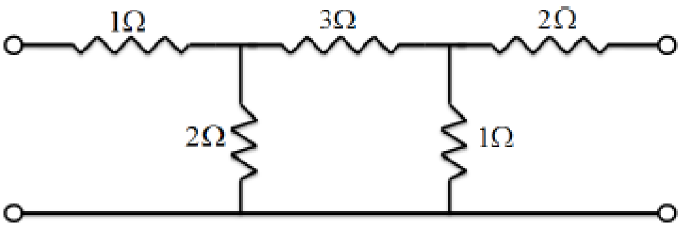
$$Z_{11} = Z_A + Z_B = 20 + 40 = 60\Omega$$

$$Z_{12} = Z_{21} = Z_B = 40 = 40\Omega$$

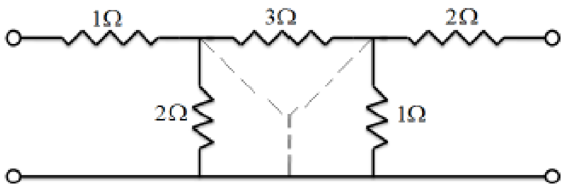
$$Z_{22} = Z_B + Z_C = 40 + 30 = 70\Omega$$

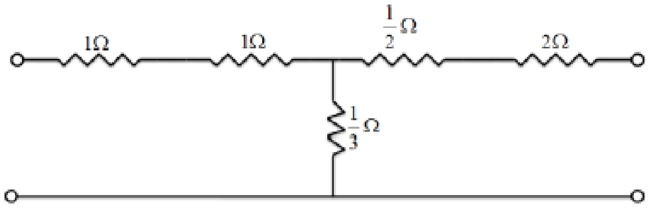
$$\therefore \qquad Z - parameter = \begin{bmatrix} 60 & 40 \\ 40 & 70 \end{bmatrix}$$

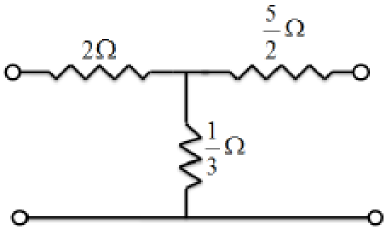
Q.Determine the Z-parameters of the network?



Solution: The circuit can be converted to a T-network, so it is easy to find the Z-parameter.







$$Z_{11} = Z_A + Z_B = 2 + \frac{1}{3} = \frac{6+1}{3} = \frac{7}{3}\Omega$$

$$Z_{12} = Z_{21} = Z_B = \frac{1}{3}\Omega$$

$$Z_{22} = Z_B + Z_C = \frac{1}{3} + \frac{5}{2} = \frac{2+15}{6} = \frac{17}{6}\Omega$$

$$Z - parameter = \begin{bmatrix} \frac{7}{3} & \frac{1}{3} \\ \frac{1}{3} & \frac{17}{6} \end{bmatrix}$$

CHAPTER -09 || FILTERS:

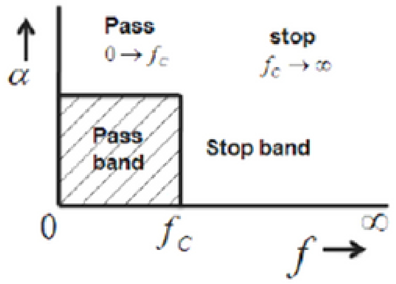
A filter is a reactive network that freely passes the desired band of frequencies while almost totally suppressing all other bands

Classification of Filters:

The **Filters** are classified into their frequency characteristics, such as,

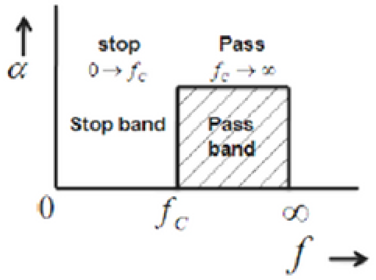
(i) **Low pass Filter (L.P.F).**
(ii) **High pass Filter (H.P.F).**
(iii) **Band pass Filter (B.P.F).**
(iv) **Band elimination Filter (B.E.P)/ Band stop filter.**

(i) Low pass Filter(L.P.F):



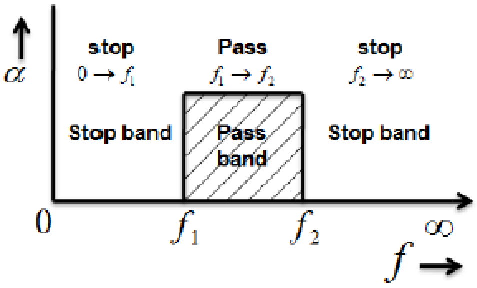
A **L.P.F** is one which passes all the frequencies without attenuation up to cut-off frequency f_C i.e. $0 \to f_C$ and suppress all the higher frequency than f_C i.e. $f_C \to \infty$.

(ii) High pass Filter(H.P.F):



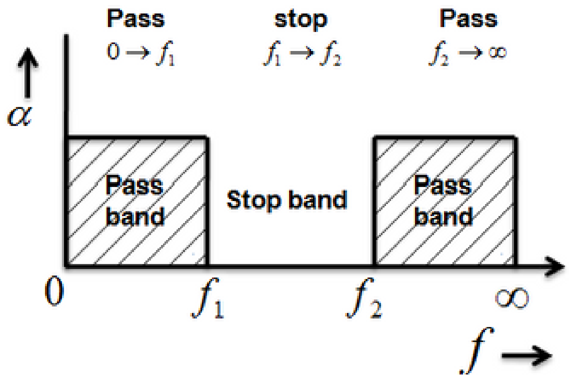
A ***H.P.F*** is one which allows all the frequency above the cut-off frequency i.e. $f_C \to \infty$ without attenuation and suppresses all lower frequency than f_C i.e. $0 \to f_C$.

(iii) Band pass Filter(B.P.F):



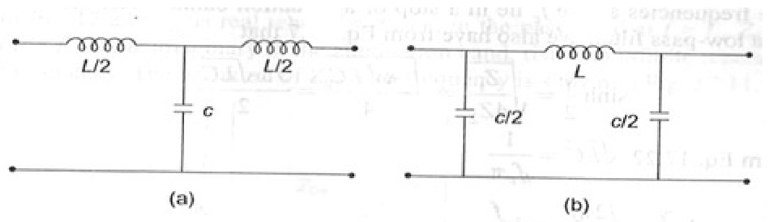
A ***B.P.F*** is one which passes the frequency between two adjacent cut-off frequencies and attenuates all other frequencies that mean ***B.P.F*** has two cut-off frequencies i.e. f_1 & f_2.

(iv) Band Elimination Filter(B.E.F):



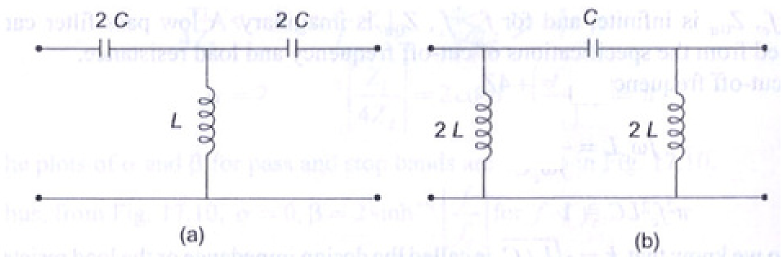
A ***B.E.F*** is one, which doesn't allow or attenuates the frequencies between two designated cut-off frequencies and allows all other frequencies.

LOW PASS FILTER



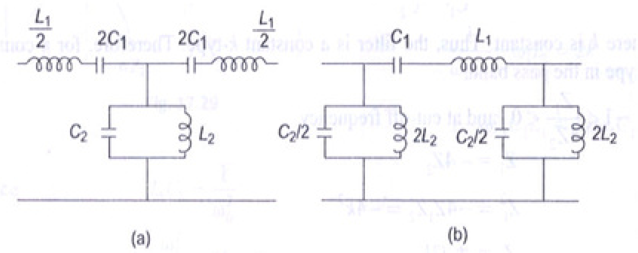
$$C = \frac{1}{\pi f_c k} \quad L = k^2 C = \frac{k}{\pi f_c}$$

HIGH PASS FILTER



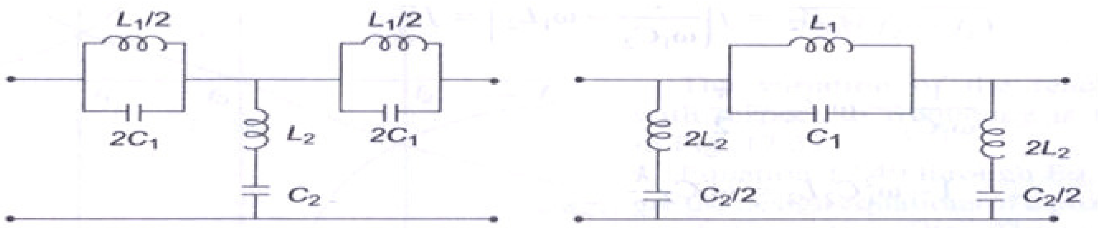
$$L = \frac{k}{4\pi f_c} \text{ and } C = \frac{1}{4\pi f_c k}$$

BAND PASS FILTER



$$C_1 = \frac{f_2 - f_1}{4\pi k f_1 f_2} \quad L_1 = \frac{k}{\pi(f_2 - f_1)} \quad L_2 = C_1 k^2 = \frac{(f_2 - f_1)k}{4\pi f_1 f_2} \quad C_2 = \frac{L_1}{k^2} = \frac{1}{\pi(f_2 - f_1)k}$$

BAND STOP FILTER



$$C_2 = \frac{1}{k\pi}\left[\frac{f_2 - f_1}{f_1 f_2}\right] \quad L_2 = \frac{k}{4\pi(f_2 - f_1)} \quad L_1 = k^2 C_2 = \frac{k}{\pi}\left(\frac{f_2 - f_1}{f_1 f_2}\right) \quad C_1 = \frac{L_2}{k^2} = \frac{1}{4\pi k(f_2 - f_1)}$$

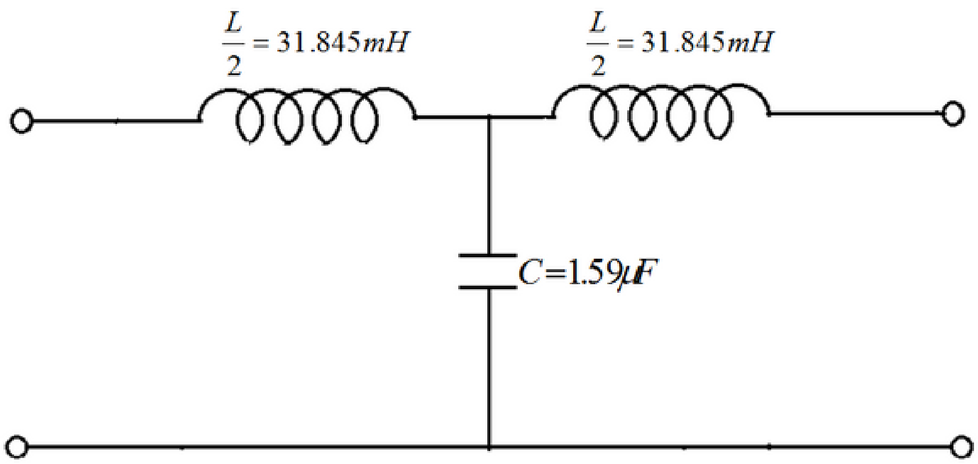
Q. Design a ***LPF*** (both T & π network) having a cut-off frequency ***1 KHz*** to operate with a terminated load resistance of ***200Ω***.

Solution:

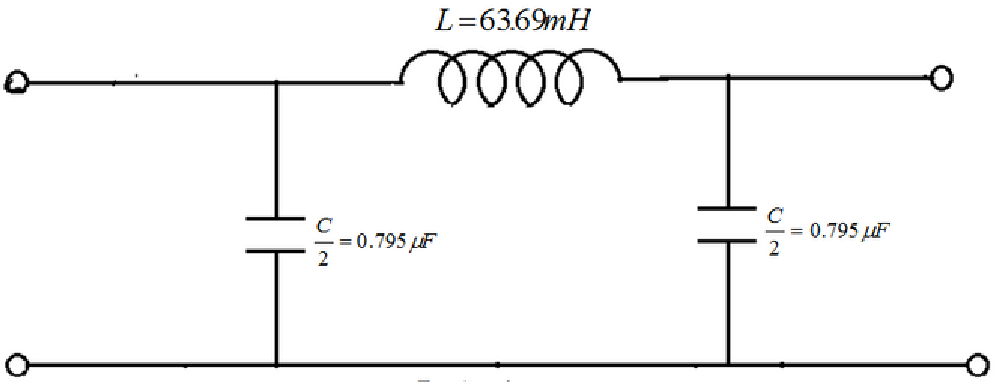
We Know that,

$$L = \frac{R_0}{\pi f_C} = \frac{200}{\pi \times 1000} = 0.06369 = 63.69mH$$

$$C = \frac{1}{\pi f_C R_0} = \frac{1}{\pi \times 1000 \times 200} = 1.59 \times 10^{-6} F = 1.59 \mu F$$



T-Network



Π-network

Q. Design a **constant-k** ***BPF*** (both T & π network) having cut-off frequency of ***3 KHz & 7.5 KHz*** and nominal characteristic impedance is ***900Ω***.

Solution: $f_{C1} = 3000Hz, f_{C2} = 7500Hz, R_0 = 900\Omega$

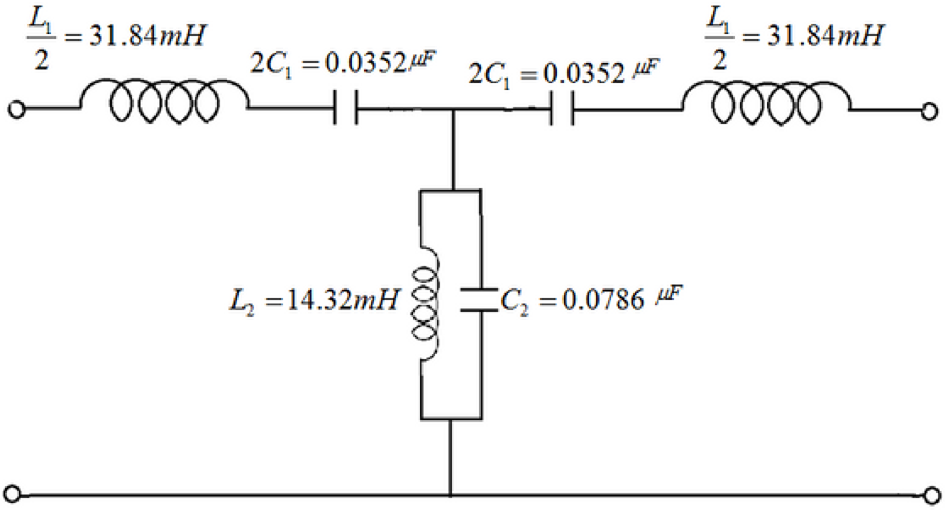
We Know that,

$$L_1 = \frac{R_0}{\pi(f_{C2} - f_{C1})} = \frac{900}{\pi(7500 - 3000)} = 0.06369H = 63.69mH$$

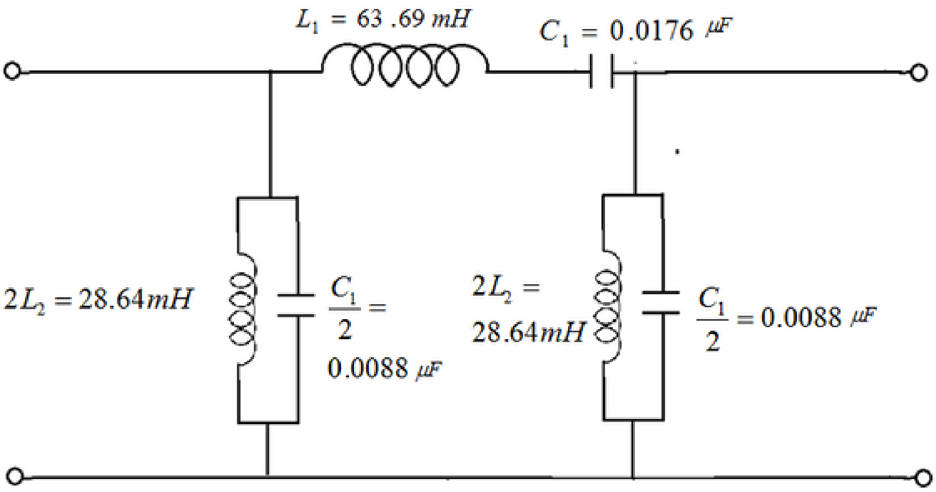
$$C_1 = \frac{f_{C2} - f_{C1}}{4\pi R_0 f_{C1} f_{C2}} = \frac{7500 - 3000}{4 \times \pi \times 900 \times 3000 \times 7500} = 0.0176\mu F$$

$$L_2 = \frac{(f_{C2} - f_{C1})R_0}{4\pi \times f_{C1} \times f_{C2}} = \frac{(7500 - 3000)}{4\pi \times 7500 \times 3000} = 14.32mH$$

$$C_2 = \frac{1}{\pi R_0 (f_{C2} - f_{C1})} = \frac{1}{\pi \times 900 \times (7500 - 3000)} = 0.078\mu F$$



T-configuration of B.P.F



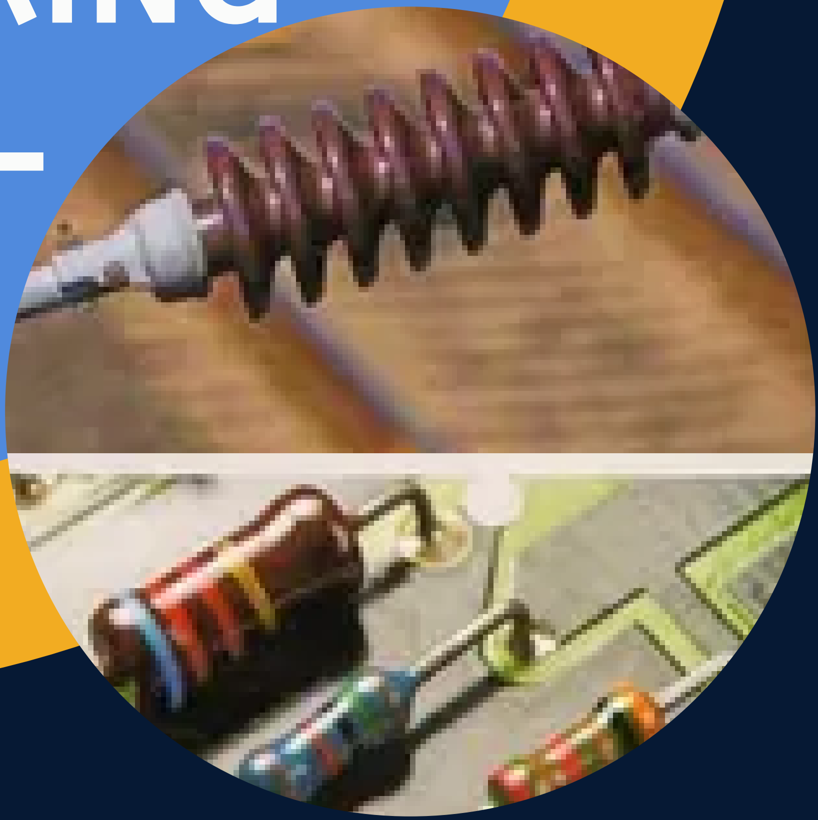
π-configuration of B.P.F

GANDHI ACADEMY OF TECHNOLOGY AND ENGINEERING

ELECTRICAL ENGINEERING MATERIAL

3rd Semester

(AS PER SCTE&VT SYLLABUS)



PREPARED BY

Mahesh Kumar Mishra

Dept. Of Electrical Engineering

INDEX

Conducting Materials

1. Introduction

An electrical engineer should posses the knowledge of the properties of materials used in electrical engineering. This knowledge helps to choose the correct materials for a given application. Hence, the materials available can be employed effectively and economically for a specific purpose.

2. Classification of Electrical Materials

- **Materials used in the electrical engineering field are classified basing on their properties and applications.**

  a. Conductor materials.

  b. Resistor materials.

  c. Insulating materials.

  d. Semiconductor materials

  e. Magnetic materials

  f. Refractory materials

  g. Structural materials.

- **Classification of Materials Based on Atomic Structure**

The materials such as gold, silver, copper and aluminium which can neither be broken into other substances nor be created are called ‘elements’. The smallest particles into which an element can be divided having the identity of the element are called ‘atoms’. These particles cannot be divided further. The atom although extremely small, has a complex internal structure of its own. This resembles the miniature solar system. An atom consists of the central core called nucleus, with electrons revolving around it as well as spinning around themselves. The nucleus contains protons and neutrons. Each proton possesses as much positive charge as an electron possesses negative charge (1.6×10^{-19} C). The number of protons inside the nucleus is equal to the number of electrons revolving around it. This number is called atomic number of the element. The neutron does not possess any charge. Therefore, the atom is electrically neutral. The mass of a proton or a neutron is 1.672 x 10^{-27} kg. which is 1850 times more than that of an electron. The mass of an electron is 9.107 x 10^{-31} kg. The electron's diameter is three times that of a proton. The weight of *protons and neutrons together* is called *atomic weight* of the element. The electrons are held in the atom by attractive force between protons and electrons which carry opposite charges. The electrons revolve in successive orbits or shells. The orbits should be visualized to be in different planes and not as they appear to be in the figure. The number of electrons that each shell can accommodate is given by $2n^2$, where n is the number of the shells counting from the innermost shell. The innermost shell (i.e. the first shell) can accommodate 2 electrons, the second shell 8, the third 18 and so on. The outermost shell in no case will contain more than 8 electrons in the first shell, 8 in the second, 8 in the third and 1 in the fourth even though the third shell can accommodate 18 electrons according to the formula.

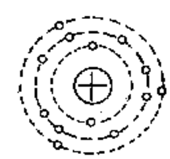


Fig 2.1

Within the shell there are sub-shells which are classified as : s, p, d, f, g, s and p and so on. There are energy levels again in these sub – shells. The sub-shell s has one energy level, p has three levels, d has five levels and so on. Not more than two electrons occupy the same energy level, one spinning in one direction and the other in the opposite direction. Thus the sub-shell

S can accommodate 1 x 2 = 2 electrons
P can accommodate 3 x 2 = 6 electrons
D can accommodate 5 x 2 = 10 electrons
F can accommodate 7 x 2 = 14 electrons
G can accommodate 9 x 2 = 18 electrons
and so on.

Shell	Possible sub-shells
1 or K	1s
2 or L	2s 2p
3 or M	3s 3p 3d
4 or N	---------
5 or O	---------
6 or P	6s 6p 6d 6f 6g 6h

According to Pauli exclusion principle, the state of any electron is defined by four Quantum numbers :

a) The shell number 1,2,3, etc. of K,L,M,N, etc.,

b) The sub-shell number s,p,d,f,g etc.

c) The orbit number in sub-shell 1s, 2s, 3s, etc., and

d) The electron spin Quantum number +1/2 and -1/2

The electrons nearer to nucleus are more firmly held than those farther from it. The energy required to pull out one electron from the first orbit is more than the energy required to pull out one electron from the second orbit and so on. That is, electrons possess a definite amount of energy, called quantum, depending upon the orbit. Hence, orbits are referred to as energy levels. The valency of an element is determined by the number of electrons it can receive or give away from its outermost sub-shell to another element in a reaction. The elements having 3 or less valency electrons, give away these electrons but elements having 5 or more valence electrons, do receive such electrons to make the total as 8, for stability. The valence electrons are very loosely held and contribute to the properties of the element. If the valence orbit contains 8 electrons, then the atom is complete and stable; if it contains less than 8, the atom is unstable and very easily gives out or receives valence electrons from the neighbor to complete its valence orbit.

3. Inter-atomic Bonds: Conductor, Semiconductor and Insulator

Inter-atomic bonding: Any solid is formed by bonding between atoms. Inter-atomic bonds are of three main types:

The first one is the ***metallic bond***. In this type, the atoms of the elements which have 1,2 or 3 valence electrons, being loosely held, give up those electron to form an electron cloud in the space of the atoms and become positive ions. The material is held together by electrostatic force between positive ions and electron cloud. The elements having small number of valence electrons are formed by this type of bonding and become ductile and have good conduction of electricity. These elements are known as ***conductors***.

The second one is called ***covalent bond***. In this bond, the atoms of the materials having 4 or more valency electrons share their electrons with neighbouring atoms as shown in Fig. The atoms of such materials behave as if they have full outer orbits. This gives full strength to the material and low electrical conductivity because no electrons are free to move. Certain materials allow valence electrons to become free by thermal energy. These elements are known as ***semiconductors***.

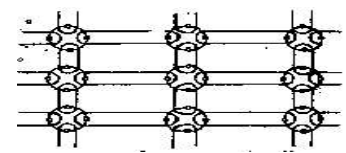


Fig 2.2

The third one is the ***ionic bond*** where the atoms of different elements transfer electrons from one to the other so that both have stable outermost orbits. For example, in sodium chloride, sodium atom gives out its one valence electron to chlorine atom and both become stable with 8 electrons in outermost orbits. At the same time, one becomes positive ion and the other negative ion. The electrostatic force between the two gives rise to the bonding. High hardness and low conductivity are typical properties of ionic bond. Therefore these materials are ***insulators***.

An atom is identified by its atomic number which indicates the number of protons in the nucleus (or the number of electrons in the orbits). For example, an oxygen atom has 8 protons and 8 neutrons in the nucleus and 8 orbital electrons. Therefore, its atomic weight is 16 and atomic number is 8.

4. Conductor Materials

- **Resistivity-**. Resistivity or specific resistance of a material may be defined as the resistance offered between the opposite faces of a metre cube of that material. The unit of resistivity is ohm metre (Ω-m). We have according to law of resistance:

  The resistance of a material (R) depends-

  - directly to its length (L)

  -inversely to the 'X'-sectional area (A)

So, R α L / A

Or $R = \rho L/A$ (where ρ is known as resistivity of material)

Therefore $\rho = A R/L$

When R= Resistance in Ohms (Ω)

L= Length in m

A= Area of cross section in m^2

ρ = resistivity or Specific resistance in Ω-m

- **Temperature Coefficient of Resistance**.

Based on temperature effect, electrical materials can be classified into two groups (i) positive temperature coefficient materials and (ii) negative temperature coefficient materials.

(i) Positive temperature coefficient means that the resistance of some of the metals and alloys increases when their temperature is raised.

(ii) Negative temperature coefficient means that the resistance of some of the materials, i.e., carbon and insulators and electrolytes, decreases when their temperature is raised.

If the resistance of a conductor is R_0 at 0^0 C, then its resistance at t^0C is given by the equation $R_t = R_0 \alpha t$ where α is the temperature coefficient of resistance at 0^0C and t is the difference in temperature.

While selecting a material for a specific purpose in electrical engineering, its electrical, mechanical and economical properties are to be considered.

5. Properties of Conductors

A. *Electrical Properties*

1. The conductivity must be good.
2. Electrical energy displayed in the form of heat must be low.
3. Resistivity must be low.
4. Temperature resistance ratio must be low.

B. ***Mechanical Properties***

1. Ductility: It has that property of a material which allows it to be drawn into a wire.
2. Solderability: The joint should have minimum contact resistance.
3. Resistance to corrosion: Should not get rusted when used in outdoors.
4. Withstand stress and strain.
5. Easy to fabricate.

C. ***Economical Factors***

1. Low Cost
2. Easily Available.

6. **Superconductor**

- Theory of super conductor.

When a piece of tin is taken and cooled down to a temperature $T_c = 3.7$ K, we find that below T_c, the tin is in a new thermodynamic state. The change that has occurred in the metal is not a change in the crystallographic structure. It is not even a ferromagnetic or anti-ferromagnetic transition. The resulting new property is that the tin has zero electrical resistance at this state. In fact, a current induced in a tin ring at temperature T_c has been observed to persist over a period of more than one year. We say that tin, in this particular condition, is a Super Conductor.

A large number of metals and alloys are superconductors, with critical temperatures T_c ranging from less 1K to 18K. Even some heavily doped semiconductors have been found to be superconductors. Historically, the first superconductor to be discovered was mercury- discovered by Kammerling Onnes in 1911.

Superconductors are by no means rare. More than 20 elements, all metals, are superconducting, and so are innumerable alloys and intermetallic compounds. Curiously enough, the best conductors like silver, copper and gold are not superconductors. Superconductivity depends on

a) electron-proton interaction, and

b) critical temperature.

- Applications.
- Superconductors can be used for the production of strong magnetic fields. Magnetic inductions in the order of 10 Wb/m^2, far above the largest value obtainable with iron-core electromagnets, have been obtained in superconducting solenoids. Other applications of superconductors are based on the effect of an applied magnetic field on the transition between normal and superconducting states. e.g. at a constant temperature below T_c, changes back and forth from normal to superconducting behavior can be affected by varying the external magnetic field, which thereby can control the current in a circuit connected to the superconductor. Thus, amplifiers, oscillators, control systems, and especially the logic and information storage functions of a large-scale computer can be provided by the controlling magnetic field exercises on superconductivity.

7. Characteristics of a Good Conductor Material

The conductor materials should have low resistivity so that the desired of a conductor material depends on the following factors :

1. Resistivity of the materials.
2. Temperature coefficient of resistance
3. Resistance against corrosion
4. Oxidation characteristics
5. Ease of soldering and welding
6. Ductility
7. Mechanical Strength
8. Flexibility and abundance
9. Durability and low cost
10. Resistance to chemicals and weather

8. **Low Resistivity Materials and their applications**

- ***Copper***

Properties :

1. Pure copper is one of the best conductors of electricity and its conductivity is highly sensitive to impurities.
2. It is reddish-brown in colour.
3. It is malleable and ductile.
4. It can be welded at red heat.
5. It is highly resistant to corrosion.
6. Melting point is $108^{4}0$C.
7. Specific gravity of copper is 8.9.
8. Electrical resistivity is 1.682 micro ohm cm.
9. Its tensile strength varies from 3 to 4.7 tonnes/cm^2.
10. It forms important alloys like bronze and gun-metal.

Uses : Wires, cables, windings of generators and transformers, overhead conductors, busbar etc.

Hard drawn (cold-drawn) copper conductor is mechanically strong with tensile strength of 40 Kg/mm^2. It is obtained by drawing cold copper bars into conductor length. It is used for overhead line conductors and busbars.

Annealed Copper (Soft Copper) Conductor. It is mechanically weak, tensile strength 20 Kg/mm^2, easily shaped into any form.

Low-resistivity Hard Copper. It is used in power cables, windings and coils as an insulated conductor. It has high flexibility and high conductivity.

- **Silver**

It is best known electrical conductor.

Properties

1. It is very costly.
2. It is not affected by weather changes.

3. It is highly ductile and malleable.

4. Its resistivity is 165 micro ohm cm.

Uses : Used in special contact, high rupturing capacity fuses, radio frequency conducting bodies, leads in valves and instruments.

- **Aluminium**

<u>Properties:</u>

1. Pure aluminium has silvery colour and lustre. It offers high resistance to corrosion. Its electrical conductivity is next to that of copper.

2. It is ductile and malleable.

3. Its electrical resistivity is 2.669 micro ohms cm at 20^0C.

4. It is good conductor of heat and electricity.

5. Its specific gravity is 2.7.

6. Its melting point is 658^0C.

7. It forms useful alloys with iron, copper, zinc and other metals.

8. It cannot be soldered or welded easily.

Uses : Overhead transmission line conductor, busbars, ACSR conductors. Well suited for cold climate.

- **Steel**.

Steel contains iron with a small percentage of carbon added to it. Iron itself is not strong but when carbon is added to it, it assumes very good mechanical properties. The tensile strength of steel is higher than that of iron. The resistivity of steel is 8-9 times higher than that of copper. Hence, steel is not generally used as conductor material. Galvanised steel wires are used as overhead telephone wires and as earth wires. Aluminium conductors are steel-reinforced to increase their tensile strength.

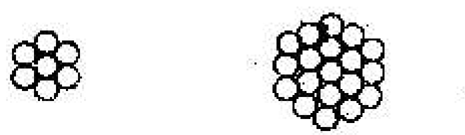
- **Bundled Conductors & Underground Cables**.

Conductor Materials for Overhead Lines : Electrical and Mechanical Properties

The function of overhead lines is to transmit electrical energy. The important properties which the line conductors must have are :

1. High electrical conductivity.
2. High tensile strength.
3. Low density.
4. Low cost.

Bundling of conductor increases the electrical and mechanical properties in comparison to the solid conductors. It is called as stranding. The number of strands in cables are 7. 19, 37, 61, 91, 127 or 169 as these conductors give the cylindrical formation.



1+6=7 Strands   1+6+12=19 Strands

Copper conductor used for transmission is hard-drawn copper.

Properties.

1. It has the best conductivity.
2. It has high current density.
3. The metal is quite homogeneous.
4. It has low specific resistance.
5. It is durable and has high scrap value.

Aluminium is next to copper to be used as a conductor.

Properties :

1. It is cheaper than copper.
2. It is lighter in weight.
3. It is second in conductivity.
4. For the same ohmic resistance, its cross-section is about 1.27 times that of copper.
5. At higher voltages, it causes lower coronal loss.

6. As the diameter of the conductor is more, it is subject to greater wind pressure due to which the swing of the conductor and sag will be greater.
7. Since the conductors are liable to swing, it requires larger cross-section.
8. As the melting point of the conductor is low, the short-circuit current will damage it.
9. Welding of aluminium is much more difficult than that of any other material.

Aluminium Conductor with Steel Reinforcement (ACSR). An aluminium conductor having a central force of galvanized steel wires is used for high-voltage transmission purposes. Reinforcement is done to increase the tensile strength of aluminium conductor. The galvanized steel core is covered by one or more strands of aluminium wires. Steel conductors used are galvanized in order to prevent rusting and electrolyte corrosion. The cross-sections of the two metals are in the ratio 1:6. For high-strength conductor, their ratio is 1:4. The steel-reinforced aluminium conductor has lower sag and longer span than the copper conductor line since it has high tensile strength. The ACSR conductor has a larger diameter than any other type of conductor of same resistance. For all calculation purposes, it is assumed that the current is passing only in the aluminium section.

Cable

Electrical and mechanical properties : Cables are most useful for low-voltage distribution in thickly populated areas. The advantages of cables are : The cable transmission is not subjected to supply interruption caused by lightning or thunderstorms, birds and other sever weather conditions. It reduces the accidents caused by breaking of the conductors. Its use does not spoil the beauty of cities.

1.10.2 Required Properties of Cables.

1. High insulation resistance.
2. Moisture and water percolated due to rain or other causes should not come in contact with conductor.
3. Low discharge current.
4. Sufficient strength for mechanical handling and cable laying.

5. Resistant to chemical action due to chemical content in earth or damages due to insects.

6. As there is not much opportunity for heat dissipation from conductor, the insulator must be capable of withstanding, without any change in qualities, the temperature within the cable.

7. It must be flexible, light and occupy less space.

8. Available in right quantity and at low rate.

Materials Used for Manufacturing Cables are Paper (impregnated), varnished cabric, vulcanized bitumen, rubber, compressed air, petroleum jelly, metal sheath (lead or lead alloy), galvanized steel or tapes for armouring and jute.

10.High Resistivity Materials and their applications

- **Tungsten**

Properties :

1. It is grayish in colour when in metallic form.
2. It has a very high melting point (3300^0C)
3. It is a very hard metal and does not become brittle at high temperature.
4. It can be drawn into very thin wires for making filaments.
5. Its resistivity is about twice that of aluminium.
6. In its thinnest form, it has very high tensile strength.
7. It oxidizes very quickly in the presence of oxygen even at a temperature of a few hundred degrees centigrade.
8. In the atmosphere of an inert gas like nitrogen or argon, or in vacuum, it will reliably work up to 2000^0C.

Uses : It is used as filaments of electric lamps and as a heater in electron tubes. It is also used in thermionic valves, radars. Grids of electronic valves, sparking and contact points.

- **Carbon**.

Carbon is mostly available as graphite which contains about 90% of carbon. Amorphous carbon is found in the form of coal, coke, charcoal, petroleum, etc.

Electrical carbon is obtained by grinding the raw carbon materials, mixing with binding agents, moulding and baking it.

Properties :

1. Carbon has very high resistivity (about 4600 micro ohm cm).
2. It has negative temperature coefficient of resistance.
3. It has a pressure-sensitive reistance material and has low surface friction.
4. The current density is 55 to 65 A/cm^2.
5. This oxidizes at about 300^0C and is very weak.
6. It has very good abrasive resistance.
7. It withstands arcing and maintains its properties at high temperature.

- **Platinum**

Properties :

1. It is a grayish-white metal.
2. It is non-corroding.
3. It is resistant to most chemicals.
4. It can be drawn into thin wires and strips.
5. Its melting point is 1775^0C.
6. Its resistivity is 10.5 micro ohm cm.
7. It is not oxidized even at high temperature.

Applications:

1. It is used as heating element in laboratory ovens and furnaces.
2. It is used as electrical contact material and as a material for grids in special-purpose vacuum tubes.
3. Platinum-rhodium thermocouple is used for measurement of temperatures up to 1600^0C.

- **Mercury**

Properties:

1. It is good conductor of heat and electricity.
2. It is a heavy silver-white metal.

3. It is the only metal which is liquid at room temperature.

4. Its electrical resistivity is 95.8 micro hom cm.

5. Oxidation takes place if heated beyond 300^0C in contact with air or oxygen.

6. It expands and contracts in regular degrees when temperature changes.

Uses : Mercury vapour lamps, mercury arc rectifiers, gas filled tubes; for making and breaking contacts; used in valves, tubes, liquid switch.

Semiconducting materials

Introduction

"A semiconductor material is one whose conductivity lies between that of a conductor and an insulator." The two most commonly used semiconductor materials are germanium and silicon.

1.3 Conductor, Insulators and Semiconductors

Any solid is formed by bonding between atoms. Inter-atomic bonds are of three main types:

The first bond is the ***metallic bond***. In this type, the atoms of the elements which have 1, 2 or 3 valence electrons, being loosely held, give up those electron to form an electron cloud in the space of the atoms and become positive ions. The material is held together by electrostatic force between positive ions and electron cloud. The elements having small number of valence electrons are formed by this type of bonding and become ductile and have good conduction of electricity. These elements are known as ***conductors***.

The second one is the ***ionic bond*** where the atoms of different elements transfer electrons from one to the other so that both have stable outermost orbits. For example, in sodium chloride, sodium atom gives out its one valence electron to chlorine atom and both become stable with 8 electrons in outermost orbits. At the same time, one becomes positive ion and the other negative ion. The electrostatic force between the two gives rise to the bonding. High hardness and low conductivity are typical properties of ionic bond. Therefore these materials are ***insulators.***

The third bond is called ***covalent bond***. In this bond, the atoms of the materials having 4 or more valence electrons share their electrons with neighbouring atoms as shown in Fig.3.1 The atoms of such materials behave as if they have full outer orbits.

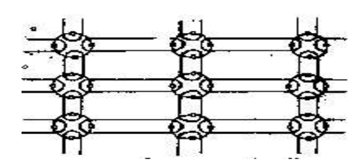


Fig 3.1

This gives full strength to the material and low electrical conductivity because no electrons are free to move. Certain materials allow valence electrons to become free by thermal energy. These elements are known as ***semiconductors***.

An atom is identified by its atomic number which indicates the number of protons in the nucleus (or the number of electrons in the orbits). For example, an oxygen atom has 8 protons and 8 neutrons in the nucleus and 8 orbital electrons. Therefore, its atomic weight is 16 and atomic number is 8.

1.5 Electron Energy and Energy Band Theory

When each atom with its neighbouring atom shares electrons in order to fill its valence ring with 8 electrons, a covalent bond is formed. Figure-3.2 shows covalent bonding.

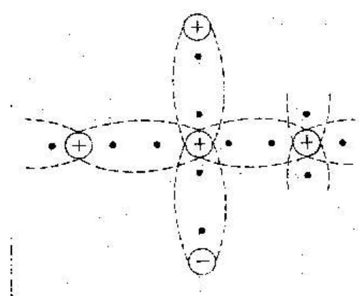


Figure -3.2 Set of covalent bond

When atoms enter into this bonding, each atom in effect has 8 valence electrons and this results in making such material a good insulator. Covalent bonding leads to the development of a poly crystal. In a poly crystal, several individual crystals are held

together imperfectly. The extra atoms are not properly locked in place. Due to impurities, there may be extra electrons which cannot lock into the covalent bond structure. Thus, a semiconductor is produced.

An impure material having three valence electrons is called trivalent bond,

e.g. Gallium, Indium and Aluminium.
An impure material having five valence electrons is called trivalent bond,
e.g. Antimony, Arsenic, Phosphorous.

Excitation of Atoms

When each electron in an atom is in its normal orbit, the atom is said to be in an unexcited state.

To move an electron further away from its nucleus requires additional energy. The additional energy can be obtained from any of the following sources: light, heat static electricity, magnetism, kinetic sources.

When the electron is in the higher energy level, the atom is said to be in an excited state. The quantum of energy, in electron volts, required to move an electron from one energy level to higher energy level varies from material to material.

When the required amount of light or heat energy is absorbed by a valence electron, it will leave the valence bond and move up to the ionization level. If it does so, it is released from the attraction forces of the nucleus. Then it is free to float between the atoms and to conduct electricity. An electron above ionization level is said to be in the conduction band and is called a free electron.

When the electron leaves the valency bond, the resulting atom is no longer neutral but has a positive charge and is called positive ion. The atom is said to be ionized.
The atom that has been ionized by the loss of an electron, does not remain so for a long time. Its positive charge will attract a nearby free electron which will give up its acquired energy. Thus, there is a constant interchange of electrons being given up and retrieved.

Energy Band Representation of lonization

In the silicon atom, K and L shells are full, but M shell contains only four electronics. According to the $2n^2$ formula, the M shell can contain 18 electronics, but the M shell in silicon is the valence shell and thus can have not more than 8 electronics.

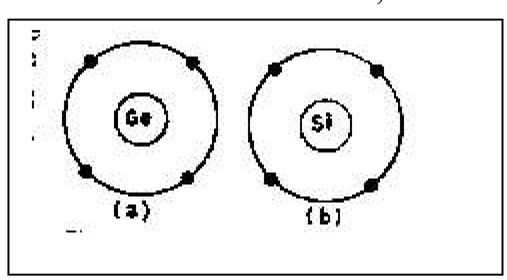


Figure -3.3 (a) simplified silicon and germanium atom

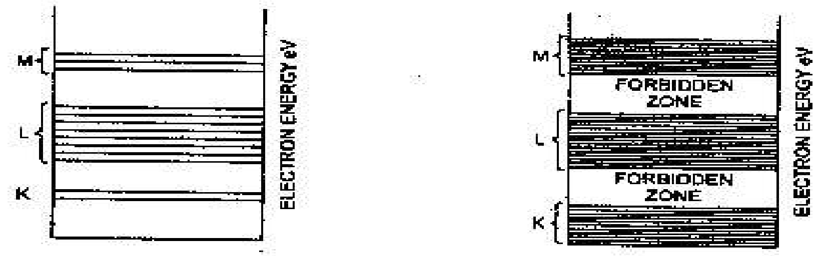


Figure -3.3 (b) Energy band representation of ionisation

In the germanium atom, the K, L and M shells are filled and the N shell is the valence shell containing 4electronics. Since only the valance electrons are important from the chemical and electrical point of view, both germanium and silicon atoms are shown in simplified from by representing only the outer most shell in Figs. 3.3 (a) and (b).

<u>Simplified Si and Ge Atoms</u>

The electrical characteristics of a semiconductor fall between those of a conductor and an insulator.

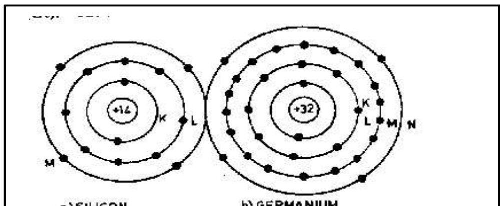


Fig- 3.4 Shell arrangement

A semiconductor has 4 electrons in its valence ring (outmost orbit). A good insulator has 8 electrons in its valence ring. The best conductor has one electron in the valence ring.

The two most widely used semiconductors are silicon (si) and germanium (Gi). Their atoms structure are shown in Figs. 3.4 (a) and (b).

<u>N-type Material.</u>

When a pentavalent impurity is added to an intrinsic material such as silicon or germanium, only four of its valence electrons lock into the covalent bond formation of atom structure. The fifth valence electron of the impurity atom is free to wander through the crystal.

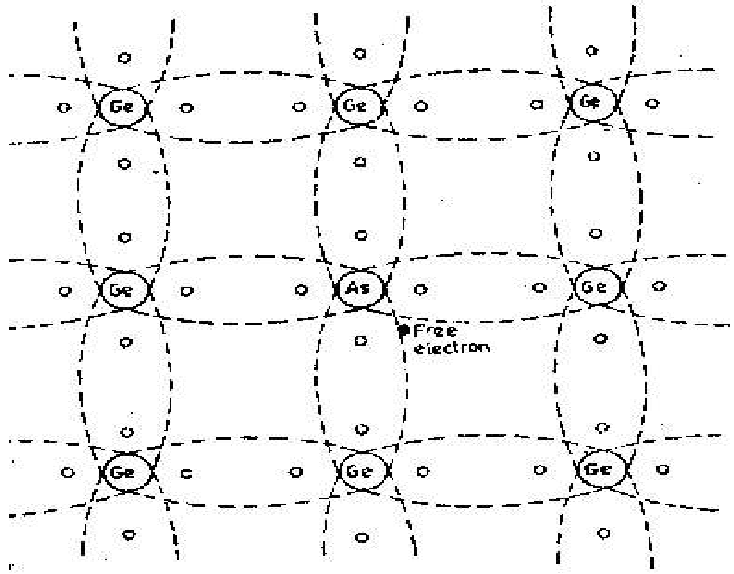


Fig- 3.5 Arsenic impurity atom provides a fifth electron

Figure 3.5 shows the addition of an atom of arsenic as an impurity. The impurity atom becomes ionized and has a positive charge when its fifth electron moves away. The positive impurity ion is not free but is firmly held in the crystal structure. The pentavalent atom donates an extra electron and is called a donor impurity. A material doped with a donor impurity has excess of electrons in its structure. It is called N-type material. The net charge of N-type material is still natural since the total number of electrons is equal to the total number of protons.

Arsenics impurity atom provides a fifth electron that cannot enter a covalent bond structure.

P-type Materials.

When a trivalent impurity is added to the intrinsic material, the two lock into a crystal structure. The impurity has three valence electrons. There is a hole in the covalent bond structure created by the lack of an electron. The hole represents an incomplete covalent bond and exhibits a positive charge. In order to complete the bond and from a stable 8-electron structure, a valence electron from a nearby atom gains sufficient energy to break loose from its bond and jumps into the hole due to its attraction. Therefore, this type of impurity is called an "acceptor". The electrons available to fill the hole and complete the bond have been release by the nearby atom whose bonds have been broken and hole created. Thus, the process will continue creating a mobility of holes. The impurity atom becomes negatively ionized as accepts an electron. The germanium or silicon atom which releases one electron become positively ionized. The net charge of the material is still neutral. The total number of electrons is equal to the total number of protons.

Semiconductors Commonly Used

The following materials are commonly used as semiconductors:

(i) Boron
(ii) Carbon
(iii) Silicon
(iv) Germanium
(v) Phosphorus
(vi) Arsenic
(vii) Antinomy
(viii) Sulphur
(ix) Selenium
(x) Tellurium
(xi) Iodine

Intrinsic Semiconductors.

If a crystal (silicon or germanium) does not contain any impure atoms (contains only one type of atoms), it is called an intrinsic material. When an electron is freed from the atom of an intrinsic material, it breaks a covalent bond and leaves behind a vacancy (called a *hole*). The free electron and the hole form an electron-hole pair. The higher the temperature, the greater the number of free electrons and holds. When a voltage is applied to an intrinsic material, it acts as a conductor.

Extrinsic Semiconductors.

Pure silicon or germanium exhibits characteristics closer to that of an insulator than a semiconductor. In order to make a material conducting, a small quantity of impurity must be added to it. The addition of impurity makes pure germanium or silicon a conductor. The process of adding impurities is called “doping”.

The extent to which the impurity has been added is called the “doping level”. When a pentavalent group provides an extra electron to the semiconductor material, the atom of the material which donates the extra electron is called a “donor atom”

When a trivalent group is added to intrinsic materials such as silicon, one covalent bond is broken, that is, a hold is created. An electron from an adjacent atom can fill the hole which is now moved to another atom. The doping atom has now one surplus negative charge and has become a negative ion. A hole is the absence of an electron and hence has a positive charge. The doping element is an “accepted”, since it takes or accepts an electron.

Majority and Minority Carriers.

In N-type material, conduction takes place through the electrons created mostly by the doping and a small number created by thermal generation.

The small number of holes created by thermal generation ove in opposite direction. In N-type material, the number of free electrons is large. These electrons are called majority carriers. Holes are in small numbers and are called minority carriers.

In p-type material, the holes are majority carriers and electrons are minority carriers.

➢ Working and Application of Semiconductors

Semiconductor materials are used in :

(i) Rectifiers

(ii) Temperature-sensitive resistors

(iii) Photoconductive and photovoltaic cells

(iv) Varistors

(v) Hall effect generators

(vi) Strain gauges

(vii) Transistors

(viii) LDR and LCD

Some of them are discussed below

<u>Germanium and Silicon Rectifiers</u>. When a P-type material and an N-type material are joined together, they form a junction called P-N junction.

When an external voltage is applied across the two material, a flow of current results if the positive and negative terminals of the voltage source are connected respectively to the ends of the P and N material. The voltage applied this way is called "forward-biasing" the P-N junction. If the applied voltage is reversed, that is, the positive of the supply voltage is connected to N side and negative of the supply is connected to the P side, there is no flow of current. This is called "reverse biasing". Thus the P-N junction offers high conductivity when forward biased and no conductivity when reverse biased. Thus, the semiconductor can be used as a rectifier. The modern P-N junction rectifiers use germanium or silicon material. Circuit diagram Fig. 3.6 a & b - below also illustrate the characteristics.

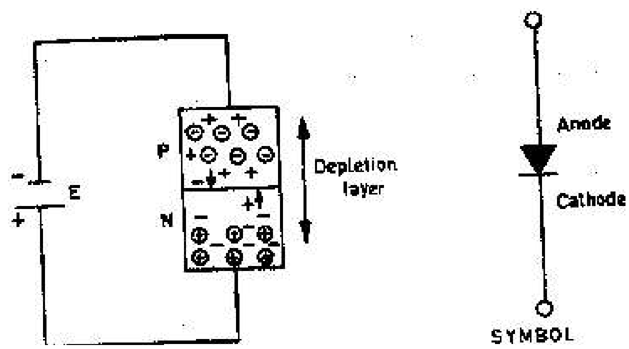


Fig-3.6 a The operation of PN junction under reverse bias

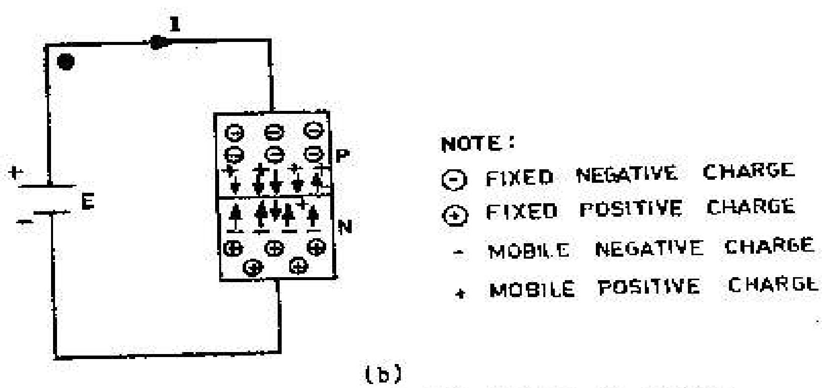


Fig- 3.6 b The operation of PN junction under forward bias

➢ **<u>Copper Oxide Rectifiers</u>**. The earliest semiconductor to be used was copper oxide. Its application was in copper oxide rectifier.

Copper oxide rectifier is a place of 99.98 % pure copper on which a film of fuprous oxide is produced by a special process. From one side of the plate, cuprous oxide is cleaned and electrode is soldered directly to the copper. The second electrode is soldered to cuprous oxide film. When a positive potential is applied to the oxide layer and negative to the copper, it corresponds to forward biasing of a P-N junction. By arranging the copper plate elements in stacks, rectifiers for use in many kinds of measuring instruments and circuits can be obtained. These rectifiers have low permissible current density. They are not used for power supply purposes.

To have a good contact with copper oxide, a lead plate is pressed against it. The two terminals of the rectifiers are the copper plate and lead plate. The oxide will be in between the plates as shown in figure- 3.7. This rectifier will allow the current to flow only from oxide to copper and will not allow flow from copper to oxide.

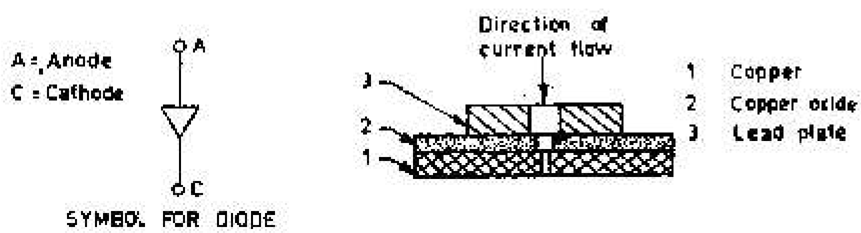


Fig- 3.7 Copper oxide rectifier

The voltage that may be applied to a single rectifier ranges between 4 and 8 V, so a number of units are connected in series for operating on high voltages. Similarly, parallel connected of the units, increases the current rating of the rectifiers, as the maximum current density in the forward direction is 0.1 to 0.15 A/cm^2 at an allowable voltage of 8 V.

The life of copper oxide rectifiers is 12 to 15 years and efficiency is 70 %.

Applications : These types of rectifiers are mostly used for meters, battery cell charging, X-ray works, measuring instruments, railway signaling, telecommunication systems, etc.

➢ **Selenium Rectifiers**.

In this type, a film of 0.5 mm. thickness is deposited on one side of the metallic back plate (iron or aluminium). By means of chemical treatment, a film of "blocking" or "barrier" layer is formed between selenium and counter electrodes.

The rectification is from back plate to selenium. The rectifier construction is as shown in figure-3.8.

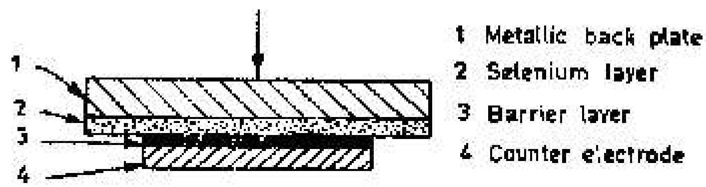


Fig-3.8 Selenium rectifier

A single unit can sustain 6 V. The normal current density is about 0.04 A per cm^2 for full wave rectification. The power efficiency is 50 to 75 %.

The units can be combined in series or in parallel, similar to that of copper oxide rectifiers to work at desired voltage or for the required current capacity.

Applications : This type of rectifiers are widely used for battery charging, telegraph and telephone circuits, control circuits, railway signaling, meters, electroplating and other works.

Such rectifiers are available in capacities of up to 50 to 100 KW.

➢ **Temperature-sensitive Elements (Thermistors)**

If the temperature of a semiconductor material is increased, that causes a decrease in its resistance. This property is used in temperature sensitive elements which are called as 'thermistors'.

The termistors are thermally sensitive material (resistors). They are made from oxides of certain metals such as copper, manganese, cobalt, iron and zinc.

Applications of thermistors: Thermistors find application in temperature measurements and control. They sense temperature variations and convert these variations into an electrical signal which is then used to control heating devices.

Thermistors are also used for measurement of radio frequency power, voltage regulation and time delay circuits.

- **Photoconductive Cells**

The resistance of semiconductor materials is low under light and increases in darkness. Phtoconductive cells can be used in applications which require the control of a certain function or event according to the colour or intensity of light.

Applications: They are used in burglar alarms, flame detectors and control for street lights.

- **Photovoltaic Cells**

Photovoltaic cells are devices that develop and emf when illuminated. They convert light energy directly into electrical energy.

Applications: The applications of photovoltaic cells are in photographic exposure meters, lighting control systems, automatic aperture control in cameral.

- **Varistors**

The resistance of semiconductors varies with the applied voltage. This property is used in devices called varistors.

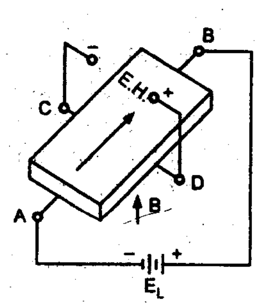
Applications. They are used in voltage stabilizers and for motor speed control.

- **Hall Effect Generators**

When a current flows through a semiconductor bar placed in a magnetic field, a voltage is developed at right angles to both current and the magnetic field. This voltage is proportional to the current and the intensity of the magnetic field. This is called the "Hall effect".

Consider the semiconductor bar shown in Fig.3.9, which has contacts on all four sides. If a voltage E_1 is applied across the two opposite sides A and B_2 a current will flow.

If the bar is placed perpendicular to magnetic field B as shown in the figure, an electrical potential E_H is generated between the other two contacts C and D. This voltage E_H is a direct measure of the magnetic field strength and can be detected with a simple voltameter.



Hall Effect Device

Fig- 3.9

Applications. The hall effect generators may be used to measure magnet is fields. It is capable of measuring magnetic field strengths that have a strength of 10^{-6} of the magnetic field of the earth.

- **Strain Gauges**

Semiconductors are sensitive to heat, voltage and magnetic field; they are also sensitive to mechanical forces. If a long thin rod of silicon is pulled from end to end, its resistance increases considerably because the mechanical force pulls each silicon atom slightly away from its adjacent atom. This increases the breadth of the forbidden energy gap, which increases the resistivity of the rod. Silicon and other semiconductors are used in strain gauges.

Applications: Strain gauges are used to find the small changes in length of solid substances or objects.

Insulating Materials

Introduction:

For safe and satisfactory operation of all electrical and electronics equipment insulator plays important role. Basically current carrying wires, surfaces need to be covered with insulating material. Let us see the structure of the material on the basis of energy band. In this type of material, the highest occupied energy band (Valence Band) is completely filled. The next higher band (Conduction Band) is quite empty.(Fig.1) The gap between these two bands is too large. When the electric field is applied across these materials, the electrons from valence band cannot reach the conduction band and conduction of electron stops. Such materials are known as insulators. Diamond is an example of this kind of material with a separation of nearly 6eV between valence band and conduction band.

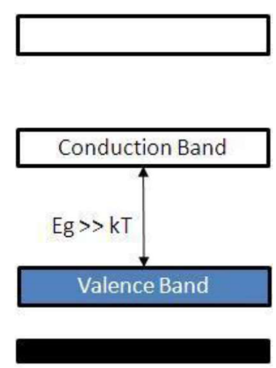


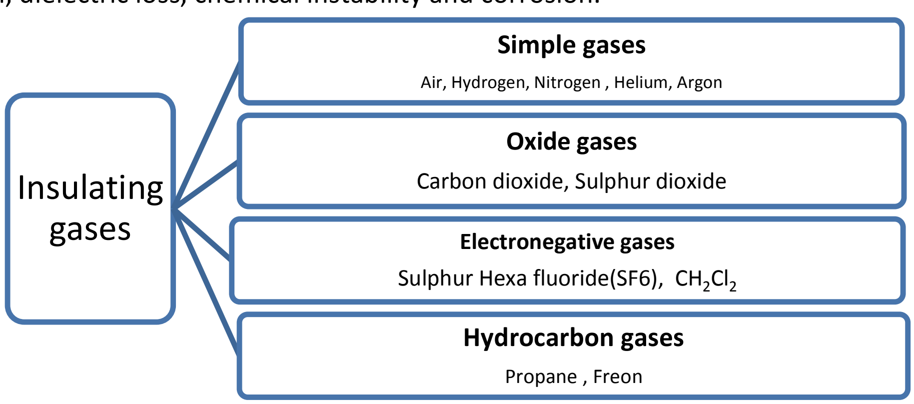
Figure 1. Energy band of INSULATOR

Insulating Materials for Electrical Engineering

The insulating materials used for various applications in electrical engineering are classified in three categories:

- Insulating gases
- Liquid insulating material
- Solid insulating material

1. ***Insulating Gases***: Many gases are used as the medium of heat transfer. All known gases are dielectric in pure form, but from electrical engineering point of view these are classified on the basis of different properties like dielectric strength, dielectric loss, chemical instability and corrosion.



2. **Liquid Insulating Material**: These materials are used for dielectric purpose to eliminate the air and other gases Insulating liquids are organic liquids used as coolant. These are categorised according to temperature range where they are used. It is used in transformers, circuit breaker, bushings, cables, capacitors etc. along with solid insulants to operate with an acceptable performance. An ideal insulating liquid material must have following properties:
    - High dielectric strength, impulse strength and volume resistivity.
    - Low dielectric dissipation factor.
    - High or low dielectric constant.(depending upon application)
    - High specific heat and thermal conductivity.
    - Excellent chemical stability and gas absorbing properties.
    - Low viscosity, density, volatility and solvent power and high flash point.
    - Good arc quenching properties.
    - Non-flammable and non toxic

Table 1 *Liquid Insulating Materials*

Type of Liquid	**Temperature Range**	**Applications**
Petroleum oils (Mineral oils)	*-50 to 110^{0}C*	All types
Askarels	*-50 to 110^{0}C*	Transformer, Capacitor, Switch gear
Silicon Liquids	*-90 to 220^{0}C*	Transformer,
Halogenated Hydro carbon	*-50 to 200^{0}C*	Electric equipment
Synthetic hydro carbon	*-50 to 110^{0}C*	Cables and Capacitors
Organic esters	*-50 to 110^{0}C*	Electronics equipments
Vegetable oil	*`-20 to 100^{0}C*	Limited application

General Properties of Insulating Material:

The suitability of an insulating material for a specific purpose use can be decided by knowing its different properties. So we have to know the exact requirement of the application and the required property hold by the insulating material. Based on uses in different applications following properties of materials are useful.

1. **Electrical Properties:** The insulating material used in electrical or electronics appliances, should be considered for following:
    - Insulation resistance
    - Dielectric constant or permittivity
    - Breakdown voltage or dielectric strength
    - Dielectric loss

1. 1 Insulation Resistance:

This is the ohmic resistance offered by an insulation coating, cover or material in an electric circuit which tends to produce a leakage current through the same with an impressed voltage across it.

Let us consider a cable of inner and outer radii r_1 and r_2, length l and resistivity of insulating material ρ. Considering a very thin layer of radial thickness *dr* at a radius *r*, the length through which the leakage current flows is *dr* and area of cross section provided to flow of current is $2\pi rl$.

Hence insulation resistance of the layer under consideration $= \frac{\rho dr}{2\pi rl}$

Insulation resistance of the cable can be determined by integrating above expression between the limits r_1 and r_2. Insulation resistance of the cable is given by,

$$R = \int_{r1}^{r2} \frac{\rho dr}{2\pi rl} = \frac{\rho}{2\pi l} \int_{r1}^{r2} \frac{dr}{r} = \frac{\rho}{2\pi l} log_e \frac{r_2}{r_1}$$

The equation states that, the resistance of the cable decreases with increase in length.

1. 2 Dielectric constant or Permittivity:

The permittivity of the insulating material varies with temperature and frequency in some cases. The materials like HCl, H_2O, CO, NH_3 have permittivity variation with change in temperature.

1. 3 Dielectric strength:

It is the maximum impressed voltage bearing capacity of insulator per unit thickness of material, up to which current does not flow through it. When current flows through the insulator is known as dielectric failure.

The dielectric strength of an insulating material decreases with the duration of time the voltage is applied, moisture, contamination, high temperature, heat ageing, mechanical stress etc. and decreases up to 10% of laboratory values.

1. 4 Dielectric loss:

Dielectric losses occur in all solid and liquid dielectric due to: a conduction current and hysteresis.

- The conduction current is due to imperfect insulating qualities of the dielectric and is calculated by the application of Ohm's law. It is in phase with the voltage and results in the power loss (I^2R) in the material, which is dissipated as heat.

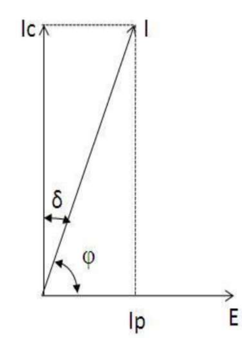


Figure 2. Plot of E against Ic

- Dielectric hysteresis is defined as the lagging of electric flux behind the electric force producing it so that under varying electric forces a dissipation of energy occurs. The energy loss due to above cause is called the dielectric hysteresis loss. The energy is dissipated as heat. This loss gives an indication of the amount of energy absorbed by the material, when subjected to AC fields.

2. Visual Properties:

An insulating material possessing two opposite properties: transparency and thermal insulation is suitable in case of reduction of energy consumption for heating and air conditioning and electrical energy savings. This is known as visual properties. Study of appearance, color and crystalline structure are the measures of this property. Glass, Aerogel hold the required visual properties. **Aerogel** is used in case of highly energy efficient windows.

3. Mechanical Properties:

Mechanical properties such as tensile strength, impact strength, toughness, hardness, elongation, flexibility, mechanical strength, abrasion resistance etc. are to be considered for choosing the insulating material.

3. 1. Mechanical Strength:

The insulating material should possess sufficient mechanical strength to respond mechanical stress. Mechanical strength is affected by following factors.

- Temperature rise: It badly affects the mechanical strength of the insulating material.
- Humidity: It is the climatic effect which affects also the mechanical strength.
- Porosity: An insulating material of high porosity will absorb more moisture and thereby affects the electrical properties as well as mechanical strength.

4. Thermal Properties:

Following thermal properties are considered for selecting insulating material of different applications.

4. 1. *Thermal stability*: The insulating material must be stable (no change in physical state) within the allowed temperatures. Certain materials like wax and plastic get soft at moderate temperatures. So the mechanical property of the material is affected. Hence the operating temperature of the material is to be noted before its use.

4. 2. *Melting point*: The insulating material should have melting point (temperature bearing capacity without being melt), above that of operating temperature.

4. 3. *Flash point*: This is an important property of insulating oils used in transformer. Flash point of a liquid insulator is that temperature at which the liquid begins to ignite.

4. 4. *Thermal conductivity*: In electrical appliances heat is generated during operation, which should be transferred to atmosphere, to maintain the operating temperature within the limit. Hence the insulators should have very low thermal conductivity

4. 5. *Thermal expansions*: Rapid and repeated load cycle on electrical appliances cause corresponding expansion and contraction of the insulators. In a result voids are created and affect the breakdown phenomenon. Thus two insulating material of different coefficient of thermal expansion should be wisely selected.

4. 6. *Heat Resistance*: The insulating material used must be able to withstand the heat produced due to continuous operation and remain stable during the operation. At the same time it should not damage the other desired properties.

5. **Chemical Properties:**

Certain chemical properties are also required to be considered for the insulating materials.

5. 1. *Chemical Resistance*: It is the ability of the insulating material to fight against corrosion in the presence of gases, water, acids and alkalis. For materials which are subjected to high voltage, high chemical resistance is also necessary.

5. 2. *Hygroscopity*: Many insulating materials are hygroscopic. Sometimes the insulation may come in direct contact with water. The porous materials are more hygroscopic than dense ones. Small amount of moisture absorbed by an insulating material affects its electrical properties drastically.

5. 3. *Moisture Permeability*: The tendency of an insulating material to pass moisture through them is known as moisture permeability. Moisture can penetrate through very small pores as the size of water molecule is very small. So this property is vital for selecting the protective coating, cable sheaths etc.

6. **Ageing:** Ageing is the long term effect of heat, chemical action and voltage application. These factors decide the natural life of insulators and hence of an electrical apparatus.

Insulating Gas: Properties and applications

Air: Air provides insulation between the over-head transmission lines. It is the best insulating material when voltages are not very high. It is also used in air capacitor, switches and various electrical equipments.

It is easily available, non-inflammable, non-explosive, small dielectric strength (nearly 3to 5 kV/m) and reliable at low voltage.

Hydrogen: It is commonly used for cooling purpose in electrical machine due to its lightness. Its high thermal conductivity helps to transmit heat from windings of high capacity alternator. Thus it reduces windage losses and increases efficiency.

Nitrogen: Nitrogen is used in place of air, to prevent oxidation due to its chemically inert property. It is generally used in transformers, gas pressure cable and capacitors.

Carbon Dioxide: Carbon dioxide is used in certain fixed type capacitor, and is used as a pre-impregnate for oil filled high voltage apparatus, such as cables and transformers. The relative permittivity of carbon oxide is 1.000985 at 0^0 C.

Sulphur Hexafluoride (SF_6): The electromagnetic gases have high dielectric strength compared to other traditional dielectric gases like nitrogen and air. The dielectric strength of SF_6 is 2.35 times more than air. The electronegative gases are non-inflammable and non-explosive. The most important gas under this category is sulphur Hexafluoride, while others are Freon gases.

SF_6 is mostly used in high voltage application and its use is most satisfactory in dielectric machines, like X-ray apparatus, Van de Graff generators, voltage stabilizers, high-voltage switch gears, gas lasers etc. SF_6 bears some special properties as follows:

- SF_6 is colourless, nontoxic and non-inflammable gas. It is the heaviest gas and has low solubility in water. The gas can be liquefied by compression. Its cooling characteristic is better than air and nitrogen.
- Under normal temperature conditions it is chemically inert and completely stable with high dielectric strength.
- This gas has very good electronegative property. Its relatively large molecules have a great affinity for free electrons, with which they combine making the gas-filled break much more resistant to dielectric breakdown.

Liquid Insulating Material: Properties and applications

Mineral oils: The operating temperature range of mineral oil is 50-110^0C. These hydrocarbon oils are used as insulating oils in transformers, circuit breakers, switch gears, capacitors etc.

In transformers, light fraction oil, such as transil oil is used to allow convection cooling. Its high flash point is 130^0C, so it is able to prevent fire hazard. Highly purified oil have a dielectric strength of 180 kv/mm and if the oil contains polar and ionizing material its dielectric loss increases.The dielectric constant is about 2.3 and therefore it is capable of dissolving only very few substances in it and produce the conducting ions. The TRANSIL oil undergoes oxidation, particularly in in the presence of catalysts such as copper, to form sludge and acids.

Light oils having Saybolt viscosity of 100 seconds at 40^0C, have been used under pressure in oil filled high voltage cables.

More viscous or tacky oils with Saybolt viscosity of 2000 seconds at 40^0C, are generally to impregnate the paper in solid type cable.

Askarels: These are non-inflammable, synthetic insulating liquids, used in temperature range of 50 – 110^0C. Chlorinated hydrocarbons are the most widely used among the askarels because of high dielectric strength, low dielectric constant (4 to 6) and small dielectric loss. They do not decompose under the influence of electric arc and have good thermal, chemical and electrical stability.

Chlorinated hydrocarbons as askarels are used as transformer fluids to reduce fire hazards. Chlorinated diphenyl, penta chloro diphenyl, trichloro diphenyl, hexa chloro diphenyl, trichloro benzene, etc., are the most widely used hydrocarbons or askarels. Askarels are generally used to impregnate a cellulose insulating material, such as paper or press board etc., for its high breakdown strength.

Silicon Fluids: It is used in the temperature range of 90-220^0C and it is clear, water like liquid. It is available in wide range of viscosity and stable in high temperature. They are non-corrosive to metal upto 200^0C and bear excellent dielectric properties in wide range of temperature. So it is used as coolants in radio pulse and aircraft transformers.

Fluorinated Liquids: These are non inflammable, chemically stable oils used in temperature range of 50-200^0C. They provide efficient heat transfer from the winding and magnetic circuits in comparison to hydrocarbon oils and used in small electric and radio devices, transformers etc. In presence of moisture electrical properties are deteriorated.

Synthetic Hydrocarbon oils: Polybutylene, Polypropylene is the example of synthetic hydrocarbon oils. They have similar dielectric strength; thermal stability and susceptibility to oxidation properties are similar as that of mineral oils. The operating temperature range is 50-110^0C. These are used in high pressure gas filled cables and dc voltage capacitors.

Organic Esters: These organic fluids are used in the temperature range of 50-110^{0}C. They have dielectric constant and very low dielectric losses. The dielectric constant ranges from 2 to 3.5. The higher range of 12.8 is obtained in tetra hydro-furyloxalate. These fluids are well suited for use in high frequency capacitors.

Vegetable Oils: These insulating liquids have temperature range of 20-100^{0}C. Drying oils are generally suitable in the formation of insulating varnishes, while non-drying oils are used as plasticisers in insulating resin compositions.

Varnish: It is the liquid form of resinous matter in oil or a volatile liquid. Hence by applying, it dries out by evaporation or chemical action to form hard, lustrous coating, which is resistant to air and water.

It is used to improve the insulation properties, mechanical strength and to reduce degradation caused by oxidation and adverse atmospheric condition.

Classification of insulating materials on the basis of structure

Classification of insulating materials is done on the basis of their physical and chemical structure.

Table 2: Classification of materials on the basis of structure of material:

Classification	Insulating Materials
Fibrous material	Wood, paper, cotton, adhesive tapes
Insulating liquids	Transformer oils, cable oils, silicone fluids
Non-resinous material	Bitumen's, wax
Glass and ceramics	Glass, porcelain etc.
Plastics	Molding powder, rubber laminations
Mineral	Mica, micanites
Gaseous	Air, H_2, N_2, Ne, CO_2, SF_6, Hg and Na vapor

Table 3: Property and uses of some common electrical engineering materials

Material	Properties	Uses
Paper and press board	Low dielectric loss, Discharge current is lower	High frequency capacitors
Cotton	Hygroscopic, Low di-electric strength, properties can be improved by impregnation	Winding of small magnetic coil, Armature winding of coil and chokes
Wood	High dielectric constant, Highly hygroscopic, dry wood can bear a voltage gradient of 10kV/inch	Terminal block, wedges of armature winding, operating rods in high voltage switch gears.
Bitumen	Hydrocarbons of jet black colour, highly soluble in mineral oil, Poor insulating property, Low hygroscopic, Acid and alkali resistant	Underground cable
Waxes	Complex organic substance, High insulating property, Low hygroscopic	As impregnated material for paper and cloth insulation, dipping medium coating on conductors.
Glass	Organic material containing oxides, silicate and borate etc. Best insulating property, High resistivity and dielectric strength.	Insulating material to form envelope for internal support in bulbs, valves, X-ray tubes, fuse casings etc.
Ceramics	Hard, Strong, Dense, Unaffected by chemical action, Stable at high temperature, Excellent dielectric properties, weak impact strength	High voltage insulation at elevated temperatures in ovens.
Asbestos	Exhibit fiber structure , Can work at high temperatures, Good tensile strengths	Capacitor dielectric, transistor, hybrid circuit substrates, Electromechanical transducers, Not useful for high voltage, Used as thermal insulators and cables in high temperature.
Rubber	Stretchable, Moisture repellant, Good insulating properties, Good corrosion resistance. Can be obtained as hard rubber, synthetic rubber, butadiene rubber, butyl rubber, chloroprene rubber and silicon rubber.	Used as protective clothing such as boots and gloves, also used as insulation covering for wires and cables. Hard rubber is used in housing for storage batteries, panel board, jacketing material.

Special Solid Insulating materials: Properties and applications

MICA: two kind of mica are used as neutral insulating material in electrical engineering. Those are Muscovite mica and Phlogophite mica.

- ***Muscovite Mica****:* The chemical composition of muscovite mica is $KAI_3SiO_3O_{10}(OH)_2$. It is translucent green, ruby, silver or brown and is strong, tough and flexible. It exhibits good corrosion resistance and is not affected by alkalis. It is used in capacitors and commutators.
- ***Phlogophite Mica:*** The chemical composition of this is, $KMg_3AlSiO_3O_{10}(OH)_2$. It possesses less flexibility. It is amber, yellow, green or grey in colour. It is more stable, but electrical properties are poorer compared to Muscovite Mica. It is used in thermal stability requirements, such as in domestic appliances like iron, hotplates etc.

Polyethylene: It is obtained by polymerization of ethylene. The polymerization is performed in the presence of catalyst at atmospheric temperature and pressure around 100^0C. To obtain heat resistance property polythene is subjected to ionizing radiation.

Polyethylene exhibits good electrical and mechanical properties, moisture resistant and not soluble in many solvents except benzene and petroleum at high temperature. The dielectric constant and power factor remains steady over a wide range of temperature.

It is used as general purpose insulation, insulations of wires and cable conductors, in high frequency cables and television circuits, jacketing material of cables. Polyurethane films are also used as dielectric material in capacitors.

Teflon: The chemical name of Teflon is Polytetra fluoro-ethylene. This is synthesized by polymerization of tetra fluro ethylene. It bears good electrical, mechanical and thermal properties. Its dielectric constant is 2 to 2.2, which does not change with time, frequency and temperature. Its insulation resistance is very high and water resistant.

It is used as dielectric materials in capacitors, covering of conductors and cables, as base material for PCBs.

Polyvinyl Chloride (PVC): It is obtained by polymerization of vinyl chloride in the presence of a catalyst at 50^0C. PVC exhibits good electrical and mechanical properties. It is hard, brittle, and non-hygroscopic and can resist flame and sun light.

PVC used as insulation material for dry batteries, jacketing material for wires and cables.

Epoxy Glass: Epoxy glass is made by bonding two or more layer of material. The layers used reinforcing glass fibers impregnated with an epoxy rasin. It is water resistant and not affected by alkalis and acids.

It is used as base material for copper-clad sheets used for PCBs, terminal port, instrument case etc.

Bakelite: It is hard, dark colored thermosetting material, which is a type of phenol formaldehyde. It is widely used for manufacture of lamp holders, switches, plug socket and bases and small panel boards.

****

Dielectric Material

Introduction:

The materials which are capable of retarding the flow of electricity or heat through them are known as dielectric or insulators. The safe handling of heat and electricity is almost impossible without use of an insulator. The material when used to prevent the loss of electrical energy and provides a safety in its operation is named as Electrical Insulating Material. The properties which are taken into consideration for an insulator are the operating temperature and breakdown voltage. However when it is used to store electrical charge, it is known as Dielectric Material.

The electrical conductivity of Dielectric material is quite low and the band gap energy is more than 3eV. This is the reason why the current cannot flow through them. The capacity of a capacitor can be increased by inserting with a dielectric material, which was discovered by Michael Faraday.

Dielectric Parameters:

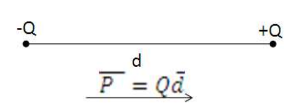
The knowledge of dielectric parameter is highly essential to choose the specific purpose dielectric for use. Those are *Dielectric constant, Dipole moment, Polarization and Polarizability.*

- ***Dielectric constant:*** The proportionality constant in the relation between the electric flux density (D) and the electric field intensity (E) is known as permittivity (ε) or dielectric constant. If the medium to which the electric field is applied is a free space (or vacuum), the proportionality constant of vacuum is ε_0 of value 8.854×10^{-12} farad.meter^{-1}. The dielectric constant of a material may be expressed as ε_r, relative to that of a vacuum by, $\epsilon_r = \frac{\epsilon}{\epsilon_0}$. So the relation of electric flux density and electric field intensity is given by,

$$D = \varepsilon_0 \varepsilon_r E$$

  Where ε_r is a dimension less quantity and is known as relative dielectric constant, which is determined by the atomic structure of the material.

- **Dipole Moment:** Two charges (Q+ and Q-) of equal magnitude but of opposite polarity, separated with distance d, constitute a dipole moment , given as: **p** = Qd
  p is the dipole moment in coulomb-meter. Dipole moment is a vector pointing from the negative charge to the positive charge and its unit is Debye (1 Debye = 3.33 × 10^{30} coulomb-metre).



Dipole Moment

- **Polarization:** The dipole moment per unit volume is called the polarization **P**. $P = \frac{p}{volume}$; where p is the dipole moment and P is the polarization in coulomb.meter^{-3}.
  Considering a parallel plate capacitor having two metal plates of area A and separated in vacuum by distance d and having a battery of voltage V connected across it. The electric field E between the plates is given by V/d volt.m^{-1} arising from the charge density ±Q on the plates. The relation between Q and E is given by, $Q = \varepsilon_0 E$.
  Q can be considered as a source of electric flux lines in the space between the plates; the density of this flux lines is the electric displacement D.

  $$D = Q = \varepsilon_0 E.$$

  Now consider that the battery is still connected and a dielectric medium is introduced to fill the space between the plates. The medium becomes polarized by the field E and dipoles appear throughout the material, lined up in the direction of the field. All dipole ends of opposite charge inside the material will cancel, but there will be an uncompensated surface charge on the plates, Positive on one plate and the negative on the other plate. These surface charges will attract and hold corresponding charges of opposite sign on the plates because the latter, unlike dipoles are able to move freely. The field in the dielectric will be still E. If the effects of some of the original surface charges have been neutralized by being bound to surface dipole ends, E can only be maintained by the flow of more charges on the battery to compensate for those, which has become bound. There is now more charge density Q' on the plates some of which is tied up and is not contributing to the field E in the dielectric. The amount of charge that is contributing to the field is the same as before and $Q' = Q + Q_B$.
  Where Q_B is the bound charge density; Q has been multiplied by a factor ε_r such that $Q' = \varepsilon_r.Q$.
  Electric field density is now given by;

  $$D = \varepsilon_0 \varepsilon_r E$$

  $$or, D = \varepsilon_0 E + Q_B$$

  The bound charge density is called polarization P. This is identical with the dipole moment per unit volume.

  The polarization may be expressed in terms of elementary dipole moments p by,

  $$P = N.p ;$$

  Where N is the number of dipoles per unit volume.

- **Polarizability:** The application of an electric field to a dielectric material causes a displacement of electric charges giving rise to the creation or reorientation of the dipoles in the material. The average dipole moment 'p' of an elementary particle may be assumed to be proportional to electric field strength E, that acts on the particle so that; $p = \alpha E$

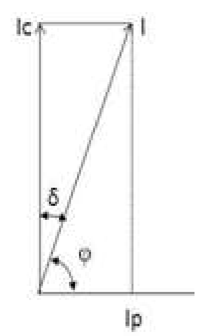
The proportionality factor α is called polirizability, measures the average dipole moment per unit field strength. The unit of the polarizability is farad.meter^2.

Mechanism of Polarization: The centre of gravity of positive charges and negative charges coincide in neutral atoms and symmetric molecules. When an electric field is applied to it, causes relative displacement of charges, leading to the creation of dipoles and hence polarization takes place. Un-symmetric arrangement of atom in a molecule results in a dipole even in the absence of an external field and in those cases the applied electric field tends to orient the dipole moments parallel to the field direction. The mechanism for forming the dipoles are categorized as (i) Electronic or Induced polarization, (ii) Ionic polarization, (c) Orientational polarization, (d) Interfacial or Space charge Polarization. Discussion of above mechanisms is restricted within the scope of the syllabus.

Dielectric Loss: The dielectric material separating the two electrodes or conductors is stressed when subject to a potential. When the potential is reversed, the stress also reversed. This change of stress involves molecularly arrangement within the dielectric. This involves the energy loss with each reversal. This is because the molecules have to overcome a certain amount of internal friction in the process of alignment. The energy expanded in the process is released as heat in the dielectric.

The loss appearing in the form of heat due to reversal of electric stresses, compelling molecular arrangement is known as dielectric loss.

When a dielectric material is subjected to an ac voltage, the leakage current I does not lead the applied voltage E by exactly 90^0. As shown in vector diagram the phase angle ϕ is always less than 90^0. The dielectric loss can be calculated as follows:



$$P = E\ I\ cos\emptyset$$

$$where\ \ \emptyset = 90^0 -\ \delta\ and\ \ I = \frac{I_c}{\cos\delta}$$

$$\therefore P = E\ \frac{I_c}{\cos\delta}\ .\cos\ 90^0 -\ \delta\ \ =\ E\ \frac{I_c}{\cos\delta}\ .sin\delta = E\ I_c tan\delta = E.\frac{E}{X_c}tan\delta$$

$$Hence, P =\ E^2\ 2\pi fc.tan\delta$$

δ is the complement angle to ϕ and is called dielectric loss angle. tanδ is the measure of dielectric loss known as dissipation factor. Good dielectric material should have very small dissipation factor yo minimize dielectric loss.

Factors affecting Dielectric loss: As observed from the equation of dielectric loss, the loss depends on the frequency and square of applied voltage. Dielectric loss increases with the presence of humidity and temperature rise.

Electrical conductivity of Dielectric and their Breakdown

The dielectric material is used in electrical and electronic circuits as insulators and as a medium in capacitors. When the applied electric field is increased, the potential difference across it also increases. A limit is reached when the dielectric ceases to work as an insulator and a spark occurs. This limiting value of the voltage is known as Breakdown Voltage, which measures the strength of dielectric.

$$Therefore, dielctric\ strength = \frac{Breakdown\ voltage}{Thickness\ of\ the\ dielectric}$$

Conduction of Gaseous dielectric: Air is the common gaseous dielectric. Cosmic rays and Ultraviolet rays cause the natural ionization in air. Since the opposite charges are equal, natural recombination takes place continuously to check further ionization of whole air.

The free charges do not go for recombination if the medium is within an Electric field. Due to application of the electric field, free charges move to their respective potential plates, causing a flow of current known as leakage current. The magnitude of current is dependent upon the applied voltage. With the increase in voltage the directed flow of electrons and ions increases as compared to random motion in low voltage. If the applied voltage is further increased, the energy of free charges becomes sufficient to force out electrons even from neutral atom. Each free electron moves at a great velocity, collides with other neutral atoms and knocks out free electron out of them. This process increases in geometric progression. The leakage current increases sharply in result to cause the breakdown of dielectric. The corresponding voltage is known as Breakdown voltage.

Conduction of Liquid dielectric: The liquid dielectric along with impurities of solid particle has more ability to conduct. The impurities get electrically charged and act as a current carrier. The fibrous impurities make the alignment of ions in a straight path for which the conductivity in liquid gets faster. In an uncontaminated liquid dielectric, such ion bridge cannot be formed. The breakdown of an uncontaminated liquid dielectric takes place due to the ionization of gases present in the liquid. The applied voltage ionizes the gas in liquid and the electric field intensity increases. It causes further ionization and ultimately the breakdown of dielectric takes place.

Conduction of solid dielectric: Electrical conductivity of solid dielectrics may be electronic, ionic or both. In electronics current flow the flow of current is due to the movement of electrons towards the positive electrodes, while ionic current flow is due to the movement of positively charged ions towards the negative electrode. The impurities also play the role of conductivity in the dielectric. At low temperatures, the conductivity of solid dielectric is due to the impurities only. At higher temperature the leakage current depends upon the contribution of free ions of the base dielectric.

Breakdown of solid dielectrics may be electro-thermal or electrical. The heat produced due to dielectric loss causes electro thermal breakdown and in effect destruction of dielectric takes place. If the dielectric is not able to radiate away the generated heat caused by dielectric loss and the applied voltage is retained for a long period the material gets melted. The electrodes get short circuited. Solid dielectric is not recoverable after its break down like liquid or gaseous dielectrics.

Properties of Dielectric Materials:

Some of the main properties of important dielectrics used in practice are given in following table:

Material	Dielectric constant	Dielectric strength (kV/mm)	tan δ	Max working temp at 0^0C	Thermal conductivity (mW/mK)	Relative density
Air	1	3	-	-	0.025	0.0013
Alcohol	2.6	-	-	-	180	0.79
Asbestos	2	2	-	400	80	3.0
Cellulose film	5.8	28	-	-	-	0.08
Cotton fabric (dry)	-	0.5	-	95	80	-
Impregnated	-	2.0	-	95	250	-
Ebonite	2.8	50	0.005	80	150	14
Glass (flint)	6.6	6	-	-	1100	4.5
Glass (crown)	4.8	6	0.02	-	600	2.2
Mica	6	40	0.02	750	600	2.8

Dry Paper	2.2	5	0.007	19	130	0.82
Impregnated paper	3.2	15	0.06	90	140	1.1
Quartz	5.7	15	0.008	1000	1000	2.4
Vulcanized Rubber	4	10	0.01	70	250	1.5
Resin	3	-	-	-	-	1.1
Fused Silica	3.6	14	_	-	-	-
Silk	-	-	-	95	60	1.2
Sulphur	4	-	0.0003	100	220	2.0
Water	7.0	-	-	-	570	1.0
Paraffin Wax	2.2	12	0.0003	35	270	0.88

Application of Dielectrics:

The most common application of dielectric is as a capacitor to store energy. Capacitors are classified according to use of dielectrics used in their manufacture.

i. Capacitors using vacuum, air or gases as dielectric.
ii. Capacitors using mineral oil as dielectric.
iii. Capacitors using a combination of solid and liquid dielectrics.
iv. Capacitors only with solid dielectrics like glass and mica etc.

****

\

MAGNETIC MATERIALS

Materials in which a state of magnetization can be induced are called magnetic materials when magnetized such materials create a magnetic field in the surrounding space.

The property of a material by virtue of which it allows itself to be magnetized is called permeability. The permeability of free space is denoted by μ_o. Its value $\mu_o = 4\pi \times 10^{-7}$.

The material permeability $\mu = \mu_o \times \mu_r$

When μ_r = Relative permeability

<u>Classification of Magnetic materials :</u>

Magnetic materials classified as :

a. Diamagnetic material
b. Para-magnetic material
c. Ferro-magnetic material

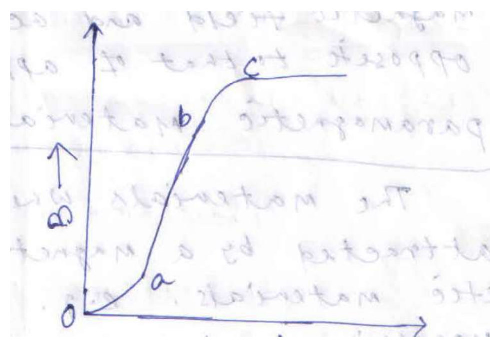
<u>DIAMAGNETIC MATERIAL :</u>

The materials which are repelled by a magnet are known as diamagnetic materials. Eg. Zinc, Mercury, lead, Sulphur, Copper, Silver. Their permeability is slightly less than one. They are slightly magnetized when placed in a strong magnetic field and act in the direction opposite to that of applied magnetic field.

PARAMAGNETIC MATERIALS :

The materials which are not strongly attracted by a magnet are known as paramagnetic materials. Eg. Aluminium, Tin, Platinum, Magnesium, Manganese, etc. Their relative permeability's is small but positive. Such materials are slightly magnetized when placed in a strong magnetic field and act in the direction of the magnetic field.

In paramagnetic materials the individual atomic dipoles are oriented in a random fashion. So the resultant magnetic field is negligible. When an external magnetic field is applied. The permanent magnetic dipoles orient themselves parallel to the applied magnetic field and give rise to a positive magnetization.



FERRO-MAGNETIC MATERIALS

The materials which are strongly attracted by a magnet are known as ferro-magnetic materials. Their permeability is very high. Eg. Iron, Nikel, Cobalt, etc. The opposing magnetic effects of electron orbital motion and electron spin do not eliminate each other in an atom of such a material.

MAGNETISIATION CURVE :

The curve drawn giving relationship between induction density 'B' and magnetizing force 'H' is known as magnetization curve or B ~ H curve.

This figure shows the general shape of B ~ H curve of magnetic material. In general it has four distinct regions oa, ab, bc and the regions beyond c. During the region oa the increase in flux density is very small, in region ab the flux density B increases almost linearly with the magnetizing force H, in region bc the increase in flux density is again small and in region beyond point c, the flux density 'B' is almost constant. The flat part of the magnetization curve corresponds to magnetic saturation of the material.

HYSTERSIS :

Hystersis is especially pronounced in materials of high residual magnetism such as hard steel. In most cases hystersis is a liability as it causes dissipation of heat, waste of energy and humming due to change in polarity and rotation of element magnets in the material.

If a magnetic substance is magnetized in a strong magnetic field it retains some portion of magnetism after the magnetic force is withdraw. The phenomenon of lagging of magnetization or induction flux density behind the magnetizing force is known as magnetic hystersis.

The losses due to hystersis is known as hystersis loss. Hystersis loss depends upon the maximum flux density 'B_m' and frequency of variation of flux is expressed as :

Hysteresis loss = $\eta B_m^{1.6} fv$ J/S or Watt.

Where η = is a constant. It is known as steinmetz hystersis coefficient

f = frequency of reversal of magnetization

B_m = Maximum flux density

V = Volume of magnetic material

EDDY CURRENT AND EDDY CURRENT LOSS :

When magnetic material is placed in alternating magnetic field, it cuts the magnetic flux. According to laws of electromagnetic induction an emf is induced. This emf causing current is known as Eddy current. The power loss due to the flow of this current is known as Eddy current loss.

Eddy current loss is proportional to the square of the frequency and the square of the thickness of the material and is inversely proportional to the resistivity of the material.

The expression for Eddy current loss is :

Eddy current loss = $KBm^2f^2t^2v^2$ Watt

Where Bm = Maximum flux density

f = Frequency of magnetic reversal

t = Thickness of lamination

v = Volume of magnetic material

CURIE POINT :

A critical temperature above which the ferro-magnetic material lose their magnetic properties is known as curie point.

SOFT AND HARD MAGNETIC MATERIALS :

All ferro-magnetic material may be divided into two types :

a. Soft-magnetic materials
b. Hard magnetic materials

Materials which have a steeply rising magnetization curve, relatively small and narrow hysteresis loop and consequently small energy losses during cyclic magnetization are called as soft magnetic materials. Soft magnetic materials are soft-iron, nickel iron alloys and soft ferrites.

Magnetic materials which have a gradually rising magnetization curve, large hysteresis loop and consequently large enrgy loss of each cycle of magnetization are called hard magnetic material. Such materials are used for making permanent magnet. The examples of hard magnetic materials are carbon-steel, tungsten-steel, cobalt steel, alnico, hard ferrites.

SOFT MAGNETIC MATERIALS :

Soft magnetic materials are used for the construction of crores for electric machines, transformers, electromagnets, reactors, relays.

In order to keep the magnetizing current and iron losses low using a low flux density. It increases the cross-sectional area of magnetic path.

The magnetic material for the core of electrical machines and transformers should have high saturation value and high permeability to keep the magnetizing current within reasonable low values. When the core is to be used for alternating magnetic fields the core should be a such material as to produce small iron losses.

PURE IRON :

Pure iron is a ferrous material with an extra-low carbon content. Eg. Low-carbon steel, electrolytic iron. The resistivity of pure iron is very low by virtue of which it gives rise to large eddy current losses when operated at high flux densities in alternating magnetic fields. Pure iron is used in many kinds of electrical apparatus and instruments as magnetic material core for electromagnets, components for relay electrical instruments.

IRON SILICON ALLOYS :

The chief alloying constituent is silicon which is added to iron in amounts from about 0.5 to 5% by weight. Iron-Silicon alloy usually known as Silicon steel. Silicon steel generally used in transformers, electrical rotating machines, reactors, electro magnets and relays.

Silicon sharply increases the electrical resistivity of iron thus decreasing the iron losses due to eddy currents. It increases the permeability at low and moderate flux densities but decreases it at high densities. Addition of silicon to iron reduce the hysteresis loss. The magneto striction effect is reduced.

Addition of silicon is valuable because it facilitates the steel making process. Alloying of low carbon steel with silicon increases the tensile strength, it reduces ductivity making steel brittle. This makes silicon alloyed steel difficult to punch and shear.

GRAIN ORIENTED SHEET STEEL :

As the ferro magnetic material have a crystal structure. So every crystal of ferro magnetic substance has a particular direction along which it offers high permeability. So it most easily magnetized. Such axes along which the crystals have high permeability and are move easily magnetized are called as easy or soft direction. Along any axis other than the easy direction, the crystal has low permeability and is therefore more difficult to magnetize. Such axes along which the crystal has low permeability are called as hard direction.

For easy magnetization the crystal directions of electrical sheet should be so oriented that their axes are paralled to the direction in which the external magnetic field is applied. This is achieved in practice by carefully controlling the rolling and annealing of silicon iron sheets. The direction of easy magnetization then lie in the direction in which the steel is rolled in the mill. Sheet steel which has been rolled such as to give easy direction to all its crystals is called ‘textured’ or grain oriented steel.

MAGNETIC ANISOTROPY :

The directional dependence of magnetic property under heading grain oriented sheet steel is known as magnetic anisotropy. It is clear that in bulk magnet a great improvement will result if the individual preferred axes are aligned parallel and along the axis of magnetization. A substantial improvement in residual magnetization and coercive force will result from parallel organization of the domain movements. The application of this technology is prevalent in the manufacture of permanent magnet.

ANNEALING :

The magnetic properties of ferro-magnetic materials are affected by strain due to mechanical working like punching, milling, grinding, machineries, etc. The magnetic properties including the correct crystal direction by heat treatment. Since mechanical stressing disturbs the crystal orientation, it is essential to perform that treatment once again after all mechanical operation have been completed.

SOFT FERRITES :

Ceramic magnet called as ferro magnetic ceramic and ferrites. Ceramic magnet are made of an iron oxide, Fe_2O_3 with one or more divalent oxides such as NiO, MnO or ZnO. These magnets have a square hysteresis loop and high resistance to demagnetization. The great advantages of territes is their high resistivity. Their resistivity's are as 10^9 Ohm-cm. Ferrites are carefully made by mixing power oxides compacting and sintering at high temperatures. High frequency transformers in television and frequency modulated receivers are almost always made with ferrite core.

HARD MAGNETIC MATERIALS :

Hard-magnetic materials are used for making permanent magnets. The properties of material required of making permanent magnets are high saturation values, high coercive force and high residual magnetism.

The hard-magnetic materials are carbon steel, tungsten steel, cobalt steel, alnico, hard ferrites.

CARBON STEEL, TUNGSTEN STEEL, COBALT STEEL :

As the soft-magnetic material have narrow hysteresis loops, so when carbon is added in a material its hysteresis loop area is increased. Although it is cheap, magnets are made from carbon steel loss their magnetic properties very fast under influence of knocks and vibrations. When materials like tungsten, chromium or cobalt are added to carbon steel, its magnetic properties are improved.

ALNICO :

It is known as Aluminium-nickel-iron-cobalt. Alnico are commercially the most important of the hard magnetic materials. Large magnets are made by special casting techniques and small one by powder metallurgy. As cobalt steel is cheaper so far this reason permanent magnets are most commonly made of Alnico.

HARD FERRITES :

Hard magnetic ferrites like BaO $(Fe_2O_3)_6$ are used for the manufacture of light weight permanent magnets due to their low specific weight.

MATERIALS FOR SPECIAL PURPOSES

Some materials used for special purposes such as fuses, solder, bimetal, storage battery plates. Those materials used for special purposes are in structural materials or protective materials.

STRUCTURAL MATERIALS :

Cast iron, steel, timber, reinforced concrete are common materials for this purpose.

Cast iron is used as materials for the frames of small and medium sized electrical machines. Steel is used in fabricated frames in large electrical machine, tanks in a transformers, fabrication of transmission towers.

Timber and reinforced concrete are used for poles in OH lines.

PROTECTIVE MATERIALS :

LEAD :

Lead is soft, heavy and bluish grey metal. It is highly resistant to many chemical action, but it can corrode by nitric acid, acetic acid, line and rotten organic substance. The electrical conductivity is 7.8% of copper. Lead is used in storage batteries and sheathing of cables. Pure lead cable sheathing are liable to fail in service due to formation of cracks formed because of vibration.

Lead alloys with tin and zinc and forms alloys which are used for solders and bearing metals.

STEEL TAPES, WIRES AND STRIPS :

Steel tapes, wires and strips are used as protective materials for mining cables, underground cable, weather proof cables.

OTHER MATERIALS :

THERMO COUPLE MATERIALS :

When two wires of different metals are joined together an emf exist across the junction. This emf is directly proportional to the temperature of the junction. When one tries to measure this emf more junctions are to be made which will give rise to emfs. When all the junctions are at the same temperature, the resultant emf will not be zero. This resultant emf is proportional to the temperature difference of the junctions and is known as thermoelectric emf.

Thermo couples are made of different materials such that copper / constantan, iron / constantan, platinum / platinum rhodium.

Thermo couples can be used for the measurement of temperature.

BIMETALS :

A bimetal is made of two metallic strips of unlike metal alloys with different coefficient of thermal expansion. At a certain temperature the strip will bend and actuate a switch or a lever of a switch. The bimetal can be heated directly or indirectly. When heated the element bends so that the metal with the greater coefficient of expansion is on the outside the are formed while that with smaller coefficient is on the inside.

Bimetallic strips are used in electrical apparatus and such as relays and regulators.

SOLDERING MATERIALS :

An alloy of two or more metals of low melting point used for base metals is known as soldering. The alloy used for joining the metals is known as solder. The solder is composed of 50% lead and 50% tin. Its melting point is 185^{o}C tensile strength is 385 kg./cm^{2} and electrical conductivity is 10% of copper.

For proper soldering flux is to be used. In soldering process the application of flux serves to removes oxides from the surface to be soldered. They deoxidize the metals at the time the soldering element is added. Solders are two types such as soft solders and hard solders soft solders are composed of lead and tin in various proportions. Hard solders may be any solder with a melting point above that of lead tin solders.

The application of soft solders is in electronic devices and hard solder in power apparatus for making permanent connection.

EYRE NO.7 FLUX :

It is an improved variety of organic flux which is used with Alca P for alumnimum cable jointing. This on decomposition at a temperature a little below the jointing temp approx 316^{o}C removes the refractory oxide from the strands of the core and makes the surface receptable to solder.

FUSE :

A fuse is a protective device, which consists of a thin wire or strip. This wire is placed with the circuit which heave to protect, so that the circuit. Current flows through it. When this current is too high the temperature of the wire or strip will increase till the wire or strip melts. So braking the circuit and interrupting the power supply.

FUSE MATERIAL :

A fuse material have following properties :

a. Low resistivity
b. Low conductivity
c. Low melting point

As lead is used as fuse material because of its low melting point. But the resistivity of lead is high, thick wires are used. For rewirable fuses alloys of tin and lead are used.

DEHYDRATING MATERIAL :

SILICA GEL :

It is an in organic chemical, a colloidal highly absorbent silica used as a de-humidifying and dehydrating agent as a catalyst carrier. Calcium chloride and silica gel are used in dehydrating breathers to remove moisture from the air entering a transformer as it breathes. Now silica gel is used for breather of a transformer. Its main advantage is that when it becomes saturated with moisture it does not restrict breathing. Silicagel when dry is blue in colour and the colour changes to pale pink as it becomes saturated with moisture.

GANDHI ACADEMY OF TECHNOLOGY AND ENGINEERING

GOLANTHARA, BERHAMPUR

LECTURE NOTES

ON

ELECTRONICS MEASUREMENT AND INSTRUMENTATIONS

For 3rd Semester

ELECTRONICS AND TELECOMMUNICATION

(As per Syllabus prescribed by SCTE&VT, Odisha)

Prepared By

Mr. Mahesh Kumar Mishra

(Lecturer in Electrical Engineering)

MEASURING INSTRUMENTS

1.1 Definition of instruments

An instrument is a device in which we can determine the magnitude or value of the quantity to be measured. The measuring quantity can be voltage, current, power and energy etc. Generally instruments are classified in to two categories.

Instrument

Absolute Instrument | **Secondary Instrument**

1.2 Absolute instrument

An absolute instrument determines the magnitude of the quantity to be measured in terms of the instrument parameter. This instrument is really used, because each time the value of the measuring quantities varies. So we have to calculate the magnitude of the measuring quantity, analytically which is time consuming. These types of instruments are suitable for laboratory use. Example: Tangent galvanometer.

1.3 Secondary instrument

This instrument determines the value of the quantity to be measured directly. Generally these instruments are calibrated by comparing with another standard secondary instrument.

Examples of such instruments are voltmeter, ammeter and wattmeter etc. Practically secondary instruments are suitable for measurement.

Secondary instruments

Indicating instruments | **Recording** | **Integrating** | **Electromechanically Indicating instruments**

1.3.1 Indicating instrument

This instrument uses a dial and pointer to determine the value of measuring quantity. The pointer indication gives the magnitude of measuring quantity.

1.3.2 Recording instrument

This type of instruments records the magnitude of the quantity to be measured continuously over a specified period of time.

1.3.3 Integrating instrument

This type of instrument gives the total amount of the quantity to be measured over a specified period of time.

1.3.4 Electromechanical indicating instrument

For satisfactory operation electromechanical indicating instrument, three forces are necessary. They are

(a) Deflecting force

(b) Controlling force

(c)Damping force

1.4 Deflecting force

When there is no input signal to the instrument, the pointer will be at its zero position. To deflect the pointer from its zero position, a force is necessary which is known as deflecting force. A system which produces the deflecting force is known as a deflecting system. Generally a deflecting system converts an electrical signal to a mechanical force.

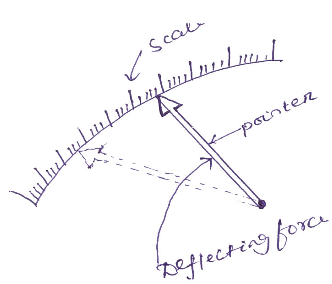


Fig. 1.1 Pointer scale

1.4.1 Magnitude effect

When a current passes through the coil (Fig.1.2), it produces a imaginary bar magnet. When a soft-iron piece is brought near this coil it is magnetized. Depending upon the current direction the poles are produced in such a way that there will be a force of attraction between the coil and the soft iron piece. This principle is used in moving iron attraction type instrument.

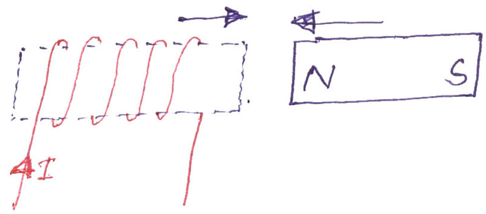


Fig. 1.2

If two soft iron pieces are place near a current carrying coil there will be a force of repulsion between the two soft iron pieces. This principle is utilized in the moving iron repulsion type instrument.

1.4.2 Force between a permanent magnet and a current carrying coil

When a current carrying coil is placed under the influence of magnetic field produced by a permanent magnet and a force is produced between them. This principle is utilized in the moving coil type instrument.

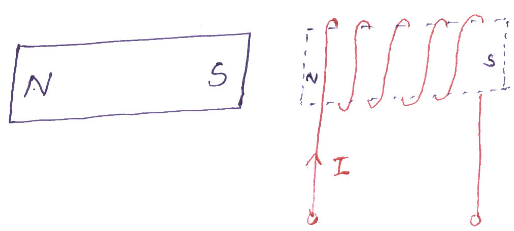


Fig. 1.3

1.4.3 Force between two current carrying coil

When two current carrying coils are placed closer to each other there will be a force of repulsion between them. If one coil is movable and other is fixed, the movable coil will move away from the fixed one. This principle is utilized in electrodynamometer type instrument.

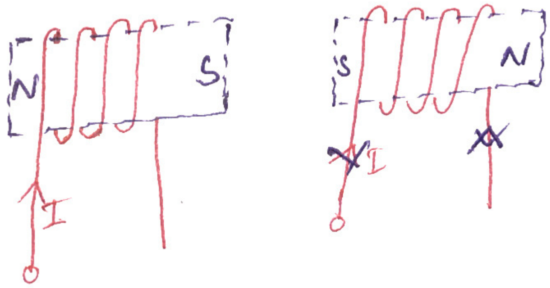


Fig. 1.4

1.5 Controlling force

To make the measurement indicated by the pointer definite (constant) a force is necessary which will be acting in the opposite direction to the deflecting force. This force is known as controlling force. A system which produces this force is known as a controlled system. When the external signal to be measured by the instrument is removed, the pointer should return back to the zero position. This is possibly due to the controlling force and the pointer will be indicating a steady value when the deflecting torque is equal to controlling torque.

$$T_d = T_c \tag{1.1}$$

1.5.1 Spring control

Two springs are attached on either end of spindle (Fig. 1.5).The spindle is placed in jewelled bearing, so that the frictional force between the pivot and spindle will be minimum. Two springs are provided in opposite direction to compensate the temperature error. The spring is made of phosphorous bronze.

When a current is supply, the pointer deflects due to rotation of the spindle. While spindle is rotate, the spring attached with the spindle will oppose the movements of the pointer. The torque produced by the spring is directly proportional to the pointer deflection θ.

$$T_C \propto \theta \tag{1.2}$$

The deflecting torque produced T_d proportional to 'I'. When $T_C = T_d$, the pointer will come to a steady position. Therefore

$$\theta \propto I \tag{1.3}$$

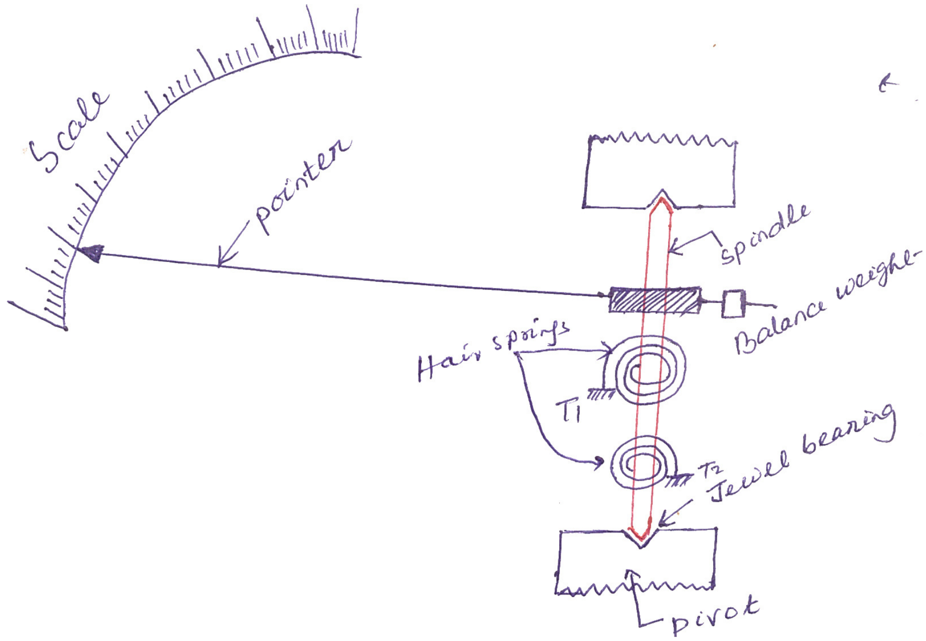


Fig. 1.5

Since, θ and I are directly proportional to the scale of such instrument which uses spring controlled is uniform.

1.6 Damping force

The deflection torque and controlling torque produced by systems are electro mechanical. Due to inertia produced by this system, the pointer oscillates about it final steady position before coming to rest. The time required to take the measurement is more. To damp out the oscillation is quickly, a damping force is necessary. This force is produced by different systems.

(a) Air friction damping
(b) Fluid friction damping
(c) Eddy current damping

1.6.1 Air friction damping

The piston is mechanically connected to a spindle through the connecting rod (Fig. 1.6). The pointer is fixed to the spindle moves over a calibrated dial. When the pointer oscillates in clockwise direction, the piston goes inside and the cylinder gets compressed. The air pushes the piston upwards and the pointer tends to move in anticlockwise direction.

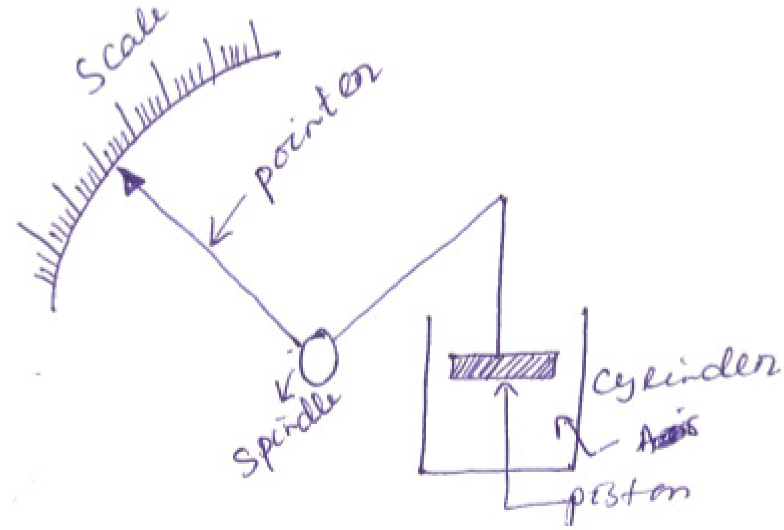


Fig. 1.6

If the pointer oscillates in anticlockwise direction the piston moves away and the pressure of the air inside cylinder gets reduced. The external pressure is more than that of the internal pressure. Therefore the piston moves down wards. The pointer tends to move in clock wise direction.

1.6.2 Eddy current damping

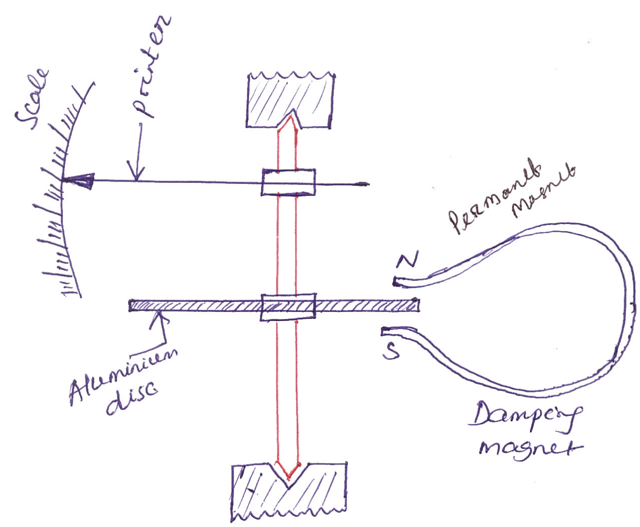


Fig. 1.6 Disc type

An aluminum circular disc is fixed to the spindle (Fig. 1.6). This disc is made to move in the magnetic field produced by a permanent magnet.

When the disc oscillates it cuts the magnetic flux produced by damping magnet. An emf is induced in the circular disc by faradays law. Eddy currents are established in the disc since it has several closed paths. By Lenz's law, the current carrying disc produced a force in a direction opposite to oscillating force. The damping force can be varied by varying the projection of the magnet over the circular disc.

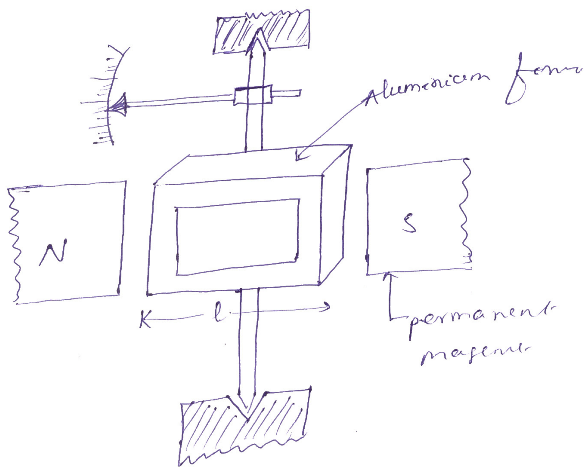


Fig. 1.6 Rectangular type

1.7 Permanent Magnet Moving Coil (PMMC) instrument

One of the most accurate type of instrument used for D.C. measurements is PMMC instrument.

<u>Construction:</u> A permanent magnet is used in this type instrument. Aluminum former is provided in the cylindrical in between two poles of the permanent magnet (Fig. 1.7). Coils are wound on the aluminum former which is connected with the spindle. This spindle is supported with jeweled bearing. Two springs are attached on either end of the spindle. The terminals of the moving coils are connected to the spring. Therefore the current flows through spring 1, moving coil and spring 2.

Damping: Eddy current damping is used. This is produced by aluminum former.

Control: Spring control is used.

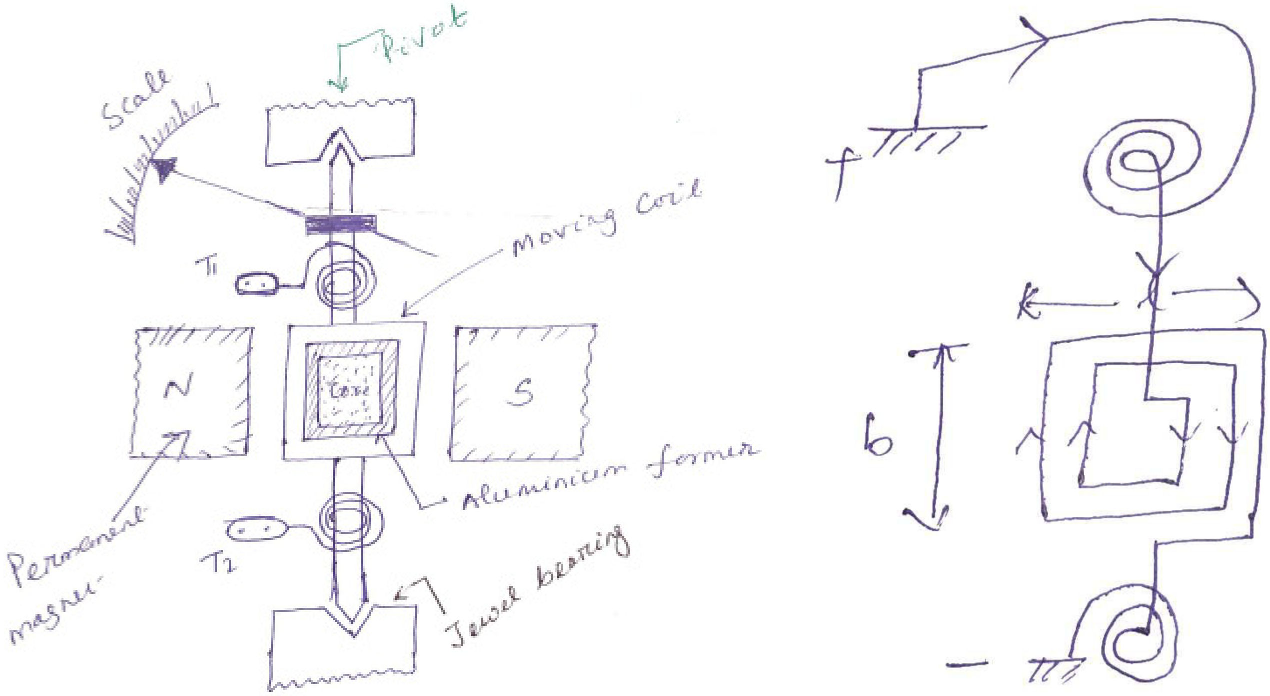


Fig. 1.7

Principle of operation

When D.C. supply is given to the moving coil, D.C. current flows through it. When the current carrying coil is kept in the magnetic field, it experiences a force. This force produces a torque and the former rotates. The pointer is attached with the spindle. When the former rotates, the pointer moves over the calibrated scale. When the polarity is reversed a torque is produced in the opposite direction. The mechanical stopper does not allow the deflection in the opposite direction. Therefore the polarity should be maintained with PMMC instrument.

If A.C. is supplied, a reversing torque is produced. This cannot produce a continuous deflection. Therefore this instrument cannot be used in A.C.

Torque developed by PMMC

Let T_d =deflecting torque

T_C = controlling torque

θ = angle of deflection

K=spring constant

b=width of the coil

l=height of the coil or length of coil

N=No. of turns

I=current

B=Flux density

A=area of the coil

The force produced in the coil is given by

$$F = BIL\sin\theta \tag{1.4}$$

When $\theta = 90^{\circ}$

For N turns, $F = NBIL$ (1.5)

Torque produced $T_d = F \times \perp_r$ distance (1.6)

$$T_d = NBIL \times b = BINA \tag{1.7}$$

$$T_d = BANI \tag{1.8}$$

$$T_d \propto I \tag{1.9}$$

Advantages

- ✓ Torque/weight is high
- ✓ Power consumption is less
- ✓ Scale is uniform
- ✓ Damping is very effective
- ✓ Since operating field is very strong, the effect of stray field is negligible
- ✓ Range of instrument can be extended

Disadvantages

- ✓ Use only for D.C.
- ✓ Cost is high
- ✓ Error is produced due to ageing effect of PMMC
- ✓ Friction and temperature error are present

1.7.1 Extension of range of PMMC instrument

Case-I: Shunt

A low shunt resistance connected in parrel with the ammeter to extent the range of current. Large current can be measured using low current rated ammeter by using a shunt.

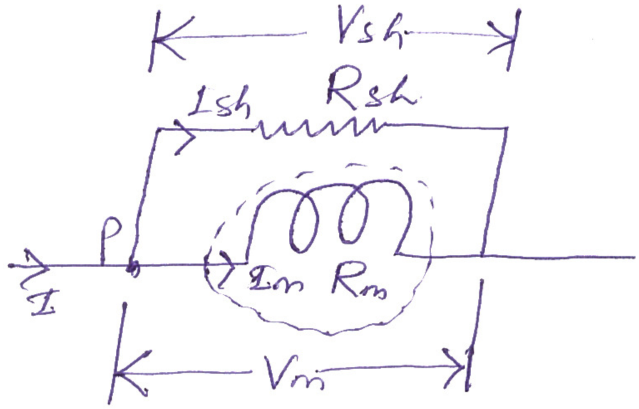


Fig. 1.8

Let R_m =Resistance of meter

R_{sh}=Resistance of shunt

I_m= Current through meter

I_{sh}=current through shunt

I= current to be measure

$\therefore V_m = V_{sh}$ (1.10)

$I_m R_m = I_{sh} R_{sh}$

$\frac{I_m}{I_{sh}} = \frac{R_{sh}}{R_m}$ (1.11)

Apply KCL at 'P' $I = I_m + I_{sh}$ (1.12)

Eq^n (1.12) ÷ by I_m

$\frac{I}{I_m} = 1 + \frac{I_{sh}}{I_m}$ (1.13)

$$\frac{I}{I_m} = 1 + \frac{R_m}{R_{sh}} \tag{1.14}$$

$$\therefore I = I_m\left(1 + \frac{R_m}{R_{sh}}\right) \tag{1.15}$$

$\left(1 + \frac{R_m}{R_{sh}}\right)$ is called multiplication factor

Shunt resistance is made of manganin. This has least thermoelectric emf. The change is resistance, due to change in temperature is negligible.

Case (II): Multiplier

A large resistance is connected in series with voltmeter is called multiplier (Fig. 1.9). A large voltage can be measured using a voltmeter of small rating with a multiplier.

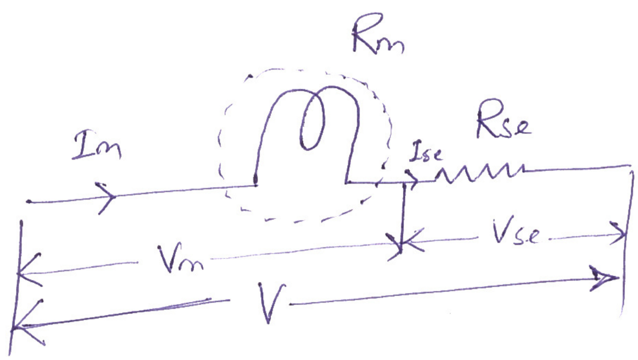


Fig. 1.9

Let R_m =resistance of meter

R_{se} =resistance of multiplier

V_m =Voltage across meter

V_{se} = Voltage across series resistance

V= voltage to be measured

$$I_m = I_{se} \tag{1.16}$$

$$\frac{V_m}{R_m} = \frac{V_{se}}{R_{se}} \tag{1.17}$$

$$\therefore \frac{V_{se}}{V_m} = \frac{R_{se}}{R_m} \tag{1.18}$$

Apply KVL, $V = V_m + V_{se}$ (1.19)

Eq[n] (1.19) $\div V_m$

$$\frac{V}{V_m} = 1 + \frac{V_{se}}{V_m} = \left(1 + \frac{R_{se}}{R_m}\right) \tag{1.20}$$

$$\therefore V = V_m\left(1 + \frac{R_{se}}{R_m}\right) \tag{1.21}$$

$\left(1 + \frac{R_{se}}{R_m}\right) \rightarrow$ Multiplication factor

1.8 Moving Iron (MI) instruments

One of the most accurate instrument used for both AC and DC measurement is moving iron instrument. There are two types of moving iron instrument.

- Attraction type
- Repulsion type

1.8.1 Attraction type M.I. instrument

Construction: The moving iron fixed to the spindle is kept near the hollow fixed coil (Fig. 1.10). The pointer and balance weight are attached to the spindle, which is supported with jeweled bearing. Here air friction damping is used.

Principle of operation

The current to be measured is passed through the fixed coil. As the current is flow through the fixed coil, a magnetic field is produced. By magnetic induction the moving iron gets magnetized. The north pole of moving coil is attracted by the south pole of fixed coil. Thus the deflecting force is produced due to force of attraction. Since the moving iron is attached with the spindle, the spindle rotates and the pointer moves over the calibrated scale. But the force of attraction depends on the current flowing through the coil.

Torque developed by M.I

Let 'θ' be the deflection corresponding to a current of 'i' amp

Let the current increases by di, the corresponding deflection is '$\theta + d\theta$'

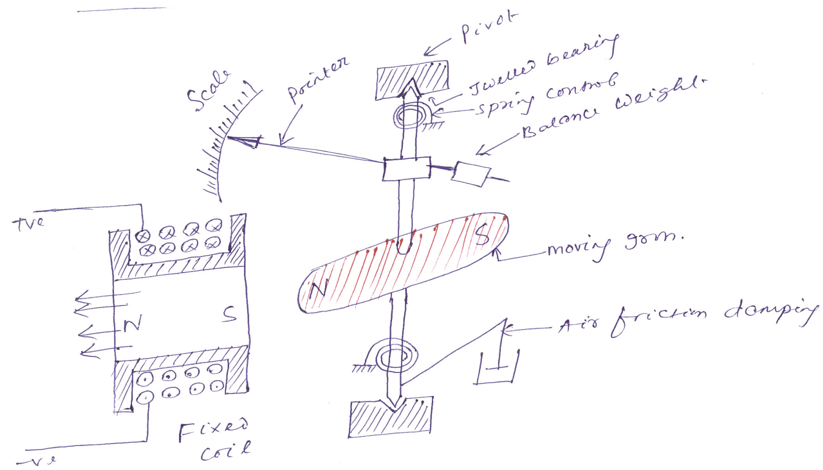


Fig. 1.10

There is change in inductance since the position of moving iron change w.r.t the fixed electromagnets.

Let the new inductance value be 'L+dL'. The current change by 'di' is dt seconds.

Let the emf induced in the coil be 'e' volt.

$$e = \frac{d}{dt}(Li) = L\frac{di}{dt} + i\frac{dL}{dt} \tag{1.22}$$

Multiplying by 'idt' in equation (1.22)

$$e \times idt = L\frac{di}{dt} \times idt + i\frac{dL}{dt} \times idt \tag{1.23}$$

$$e \times idt = Lidi + i^2 dL \tag{1.24}$$

Eq^n (1.24) gives the energy is used in to two forms. Part of energy is stored in the inductance. Remaining energy is converted in to mechanical energy which produces deflection.

Fig. 1.11

Change in energy stored=Final energy-initial energy stored

$$=\frac{1}{2}(L+dL)(i+di)^2-\frac{1}{2}Li^2$$

$$=\frac{1}{2}\{(L+dL)(i^2+di^2+2idi)-Li^2\}$$

$$=\frac{1}{2}\{(L+dL)(i^2+2idi)-Li^2\}$$

$$=\frac{1}{2}\{Li^2+2Lidi+i^2dL+2ididL-Li^2\}$$

$$=\frac{1}{2}\{2Lidi+i^2dL\}$$

$$=Lidi+\frac{1}{2}i^2dL \tag{1.25}$$

Mechanical work to move the pointer by $d\theta$

$$=T_d d\theta \tag{1.26}$$

By law of conservation of energy,
Electrical energy supplied=Increase in stored energy+ mechanical work done.

Input energy= Energy stored + Mechanical energy

$$Lidi+i^2dL=Lidi+\frac{1}{2}i^2dL+T_d d\theta \tag{1.27}$$

$$\frac{1}{2}i^2dL=T_d d\theta \tag{1.28}$$

$$T_d=\frac{1}{2}i^2\frac{dL}{d\theta} \tag{1.29}$$

At steady state condition $T_d=T_C$

$$\frac{1}{2}i^2\frac{dL}{d\theta}=K\theta \tag{1.30}$$

$$\theta=\frac{1}{2K}i^2\frac{dL}{d\theta} \tag{1.31}$$

$$\theta\propto i^2 \tag{1.32}$$

When the instruments measure AC, $\theta\propto i^2{}_{rms}$

Scale of the instrument is non uniform.

Advantages

- ✓ MI can be used in AC and DC
- ✓ It is cheap
- ✓ Supply is given to a fixed coil, not in moving coil.
- ✓ Simple construction
- ✓ Less friction error.

Disadvantages

- ✓ It suffers from eddy current and hysteresis error
- ✓ Scale is not uniform
- ✓ It consumed more power
- ✓ Calibration is different for AC and DC operation

1.8.2 Repulsion type moving iron instrument

Construction: The repulsion type instrument has a hollow fixed iron attached to it (Fig. 1.12). The moving iron is connected to the spindle. The pointer is also attached to the spindle in supported with jeweled bearing.

Principle of operation: When the current flows through the coil, a magnetic field is produced by it. So both fixed iron and moving iron are magnetized with the same polarity, since they are kept in the same magnetic field. Similar poles of fixed and moving iron get repelled. Thus the deflecting torque is produced due to magnetic repulsion. Since moving iron is attached to spindle, the spindle will move. So that pointer moves over the calibrated scale.

Damping: Air friction damping is used to reduce the oscillation.

Control: Spring control is used.

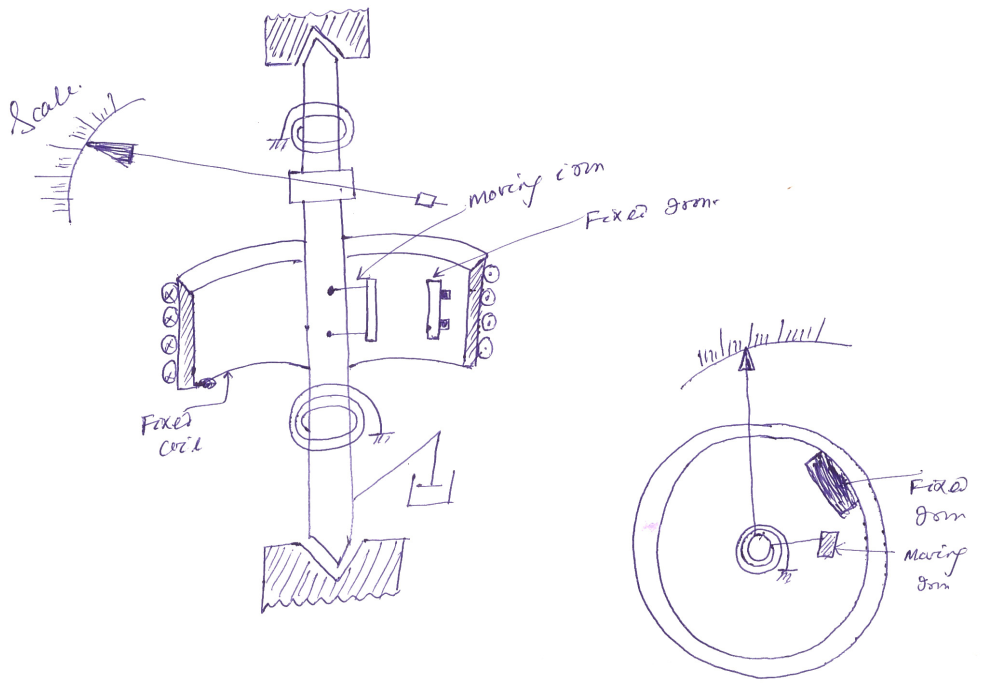


Fig. 1.12

1.9 Dynamometer (or) Electromagnetic moving coil instrument (EMMC)

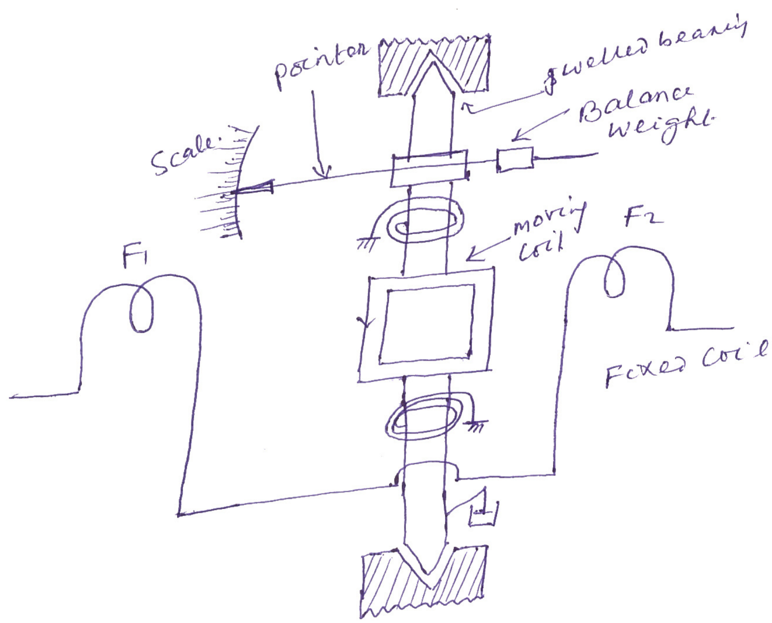


Fig. 1.13

This instrument can be used for the measurement of voltage, current and power. The difference between the PMMC and dynamometer type instrument is that the permanent magnet is replaced by an electromagnet.

Construction: A fixed coil is divided in to two equal half. The moving coil is placed between the two half of the fixed coil. Both the fixed and moving coils are air cored. So that the hysteresis effect will be zero. The pointer is attached with the spindle. In a non metallic former the moving coil is wounded.

Control: Spring control is used.

Damping: Air friction damping is used.

Principle of operation:

When the current flows through the fixed coil, it produced a magnetic field, whose flux density is proportional to the current through the fixed coil. The moving coil is kept in between the fixed coil. When the current passes through the moving coil, a magnetic field is produced by this coil.

The magnetic poles are produced in such a way that the torque produced on the moving coil deflects the pointer over the calibrated scale. This instrument works on AC and DC. When AC voltage is applied, alternating current flows through the fixed coil and moving coil. When the current in the fixed coil reverses, the current in the moving coil also reverses. Torque remains in the same direction. Since the current i_1 and i_2 reverse simultaneously. This is because the fixed and moving coils are either connected in series or parallel.

Torque developed by EMMC

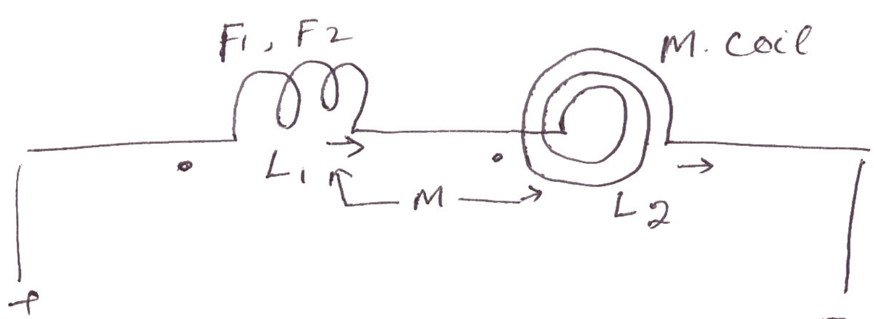


Fig. 1.14

Let

L_1=Self inductance of fixed coil

L_2= Self inductance of moving coil

M=mutual inductance between fixed coil and moving coil

i_1=current through fixed coil

i_2=current through moving coil

Total inductance of system,

$$L_{total} = L_1 + L_2 + 2M \tag{1.33}$$

But we know that in case of M.I

$$T_d = \frac{1}{2} i^2 \frac{d(L)}{d\theta} \tag{1.34}$$

$$T_d = \frac{1}{2} i^2 \frac{d}{d\theta}(L_1 + L_2 + 2M) \tag{1.35}$$

The value of L_1 and L_2 are independent of 'θ' but 'M' varies with θ

$$T_d = \frac{1}{2} i^2 \times 2 \frac{dM}{d\theta} \tag{1.36}$$

$$T_d = i^2 \frac{dM}{d\theta} \tag{1.37}$$

If the coils are not connected in series $i_1 \neq i_2$

$$\therefore T_d = i_1 i_2 \frac{dM}{d\theta} \tag{1.38}$$

$$T_C = T_d \tag{1.39}$$

$$\therefore \theta = \frac{i_1 i_2}{K} \frac{dM}{d\theta} \tag{1.40}$$

Hence the deflection of pointer is proportional to the current passing through fixed coil and moving coil.

1.9.1 Extension of EMMC instrument

Case-I Ammeter connection

Fixed coil and moving coil are connected in parallel for ammeter connection. The coils are designed such that the resistance of each branch is same.

Therefore

$I_1 = I_2 = I$

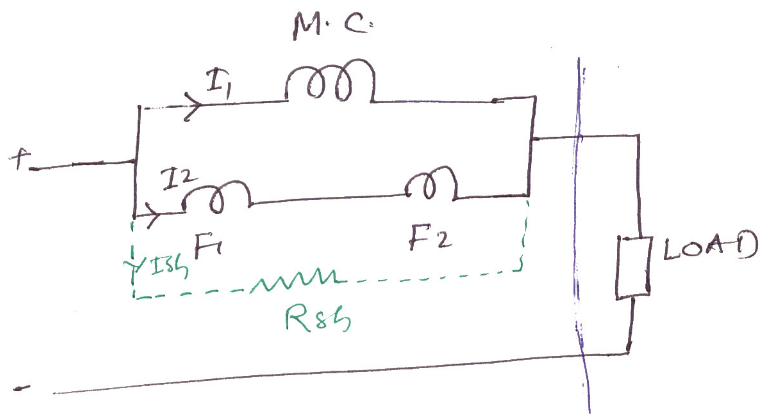


Fig. 1.15

To extend the range of current a shunt may be connected in parallel with the meter. The value R_{sh} is designed such that equal current flows through moving coil and fixed coil.

$$\therefore T_d = I_1 I_2 \frac{dM}{d\theta} \tag{1.41}$$

$$\text{Or } \therefore T_d = I^2 \frac{dM}{d\theta} \tag{1.42}$$

$$T_C = K\theta \tag{1.43}$$

$$\theta = \frac{I^2}{K}\frac{dM}{d\theta} \tag{1.44}$$

$$\therefore \theta \propto I^2 \text{ (Scale is not uniform)} \tag{1.45}$$

Case-II Voltmeter connection

Fixed coil and moving coil are connected in series for voltmeter connection. A multiplier may be connected in series to extent the range of voltmeter.

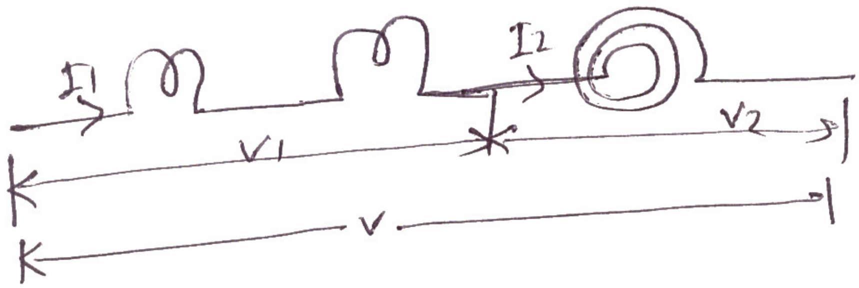


Fig. 1.16

$$I_1 = \frac{V_1}{Z_1}, I_2 = \frac{V_2}{Z_2} \tag{1.46}$$

$$T_d = \frac{V_1}{Z_1} \times \frac{V_2}{Z_2} \times \frac{dM}{d\theta} \tag{1.47}$$

$$T_d = \frac{K_1 V}{Z_1} \times \frac{K_2 V}{Z_2} \times \frac{dM}{d\theta} \tag{1.48}$$

$$T_d = \frac{KV^2}{Z_1 Z_2} \times \frac{dM}{d\theta} \tag{1.49}$$

$$T_d \propto V^2 \tag{1.50}$$

$$\therefore \theta \propto V^2 \quad \text{(Scale in not uniform)} \tag{1.51}$$

Case-III As w**attmeter**

When the two coils are connected to parallel, the instrument can be used as a wattmeter. Fixed coil is connected in series with the load. Moving coil is connected in parallel with the load. The moving coil is known as voltage coil or pressure coil and fixed coil is known as current coil.

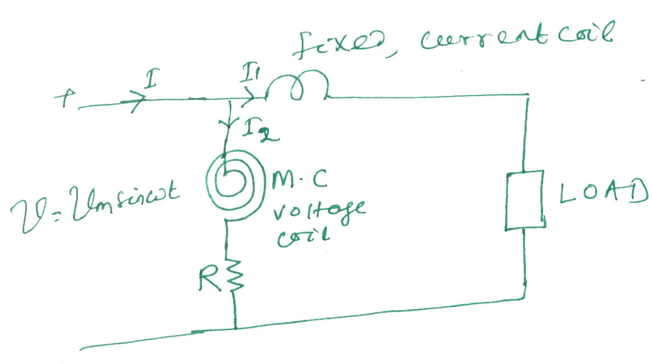


Fig. 1.17

Assume that the supply voltage is sinusoidal. If the impedance of the coil is neglected in comparison with the resistance ‘R’. The current,

$$I_2 = \frac{v_m \sin wt}{R} \tag{1.52}$$

Let the phase difference between the currents I_1 and I_2 is ϕ

$$I_1 = I_m \sin(wt - \phi) \tag{1.53}$$

$$T_d = I_1 I_2 \frac{dM}{d\theta} \tag{1.54}$$

$$T_d = I_m \sin(wt - \phi) \times \frac{V_m \sin wt}{R} \frac{dM}{d\theta} \tag{1.55}$$

$$T_d = \frac{1}{R}(I_m V_m \sin wt \sin(wt - \phi)) \frac{dM}{d\theta} \tag{1.56}$$

$$T_d = \frac{1}{R} I_m V_m \sin wt.\sin(wt - \phi) \frac{dM}{d\theta} \tag{1.57}$$

The average deflecting torque

$$(T_d)_{avg} = \frac{1}{2\Pi} \int_0^{2\Pi} T_d \times d(wt) \tag{1.58}$$

$$(T_d)_{avg} = \frac{1}{2\Pi} \int_0^{2\Pi} \frac{1}{R} \times I_m V_m \sin wt.\sin(wt - \phi) \frac{dM}{d\theta} \times d(wt) \tag{1.59}$$

$$(T_d)_{avg} = \frac{V_m I_m}{2 \times 2\Pi} \times \frac{1}{R} \times \frac{dM}{d\theta} \left[\int \{\cos\phi - \cos(2wt - \phi)\} dwt \right] \tag{1.60}$$

$$(T_d)_{avg} = \frac{V_m I_m}{4\Pi R} \times \frac{dM}{d\theta} \left[\int_0^{2\Pi} \cos\phi.dwt - \int_0^{2\Pi} \cos(2wt - \phi).dwt \right] \tag{1.61}$$

$$(T_d)_{avg} = \frac{V_m I_m}{4\Pi R} \times \frac{dM}{d\theta} \left[\cos\phi [wt]_0^{2\Pi} \right] \tag{1.62}$$

$$(T_d)_{avg} = \frac{V_m I_m}{4\Pi R} \times \frac{dM}{d\theta} [\cos\phi (2\Pi - 0)] \tag{1.63}$$

$$(T_d)_{avg} = \frac{V_m I_m}{2} \times \frac{1}{R} \times \frac{dM}{d\theta} \times \cos\phi \tag{1.64}$$

$$(T_d)_{avg} = V_{rms} \times I_{rms} \times \cos\phi \times \frac{1}{R} \times \frac{dM}{d\theta} \tag{1.65}$$

$$(T_d)_{avg} \propto KVI\cos\phi \qquad (1.66)$$

$$T_C \propto \theta \qquad (1.67)$$

$$\theta \propto KVI\cos\phi \qquad (1.68)$$

$$\theta \propto VI\cos\phi \qquad (1.69)$$

Advantages

- ✓ It can be used for voltmeter, ammeter and wattmeter
- ✓ Hysteresis error is nill
- ✓ Eddy current error is nill
- ✓ Damping is effective
- ✓ It can be measure correctively and accurately the rms value of the voltage

Disadvantages

- ✓ Scale is not uniform
- ✓ Power consumption is high(because of high resistance)
- ✓ Cost is more
- ✓ Error is produced due to frequency, temperature and stray field.
- ✓ Torque/weight is low.(Because field strength is very low)

Errors in PMMC

- ✓ The permanent magnet produced error due to ageing effect. By heat treatment, this error can be eliminated.
- ✓ The spring produces error due to ageing effect. By heat treating the spring the error can be eliminated.
- ✓ When the temperature changes, the resistance of the coil vary and the spring also produces error in deflection. This error can be minimized by using a spring whose temperature co-efficient is very low.

1.10 Difference between attraction and repulsion type instrument

An attraction type instrument will usually have a lower inductance, compare to repulsion type instrument. But in other hand, repulsion type instruments are more suitable for economical production in manufacture and nearly uniform scale is more easily obtained. They are therefore much more common than attraction type.

1.11 Characteristics of meter

1.11.1 Full scale deflection current(I_{FSD})

The current required to bring the pointer to full-scale or extreme right side of the instrument is called full scale deflection current. It must be as small as possible. Typical value is between 2 μ A to 30mA.

1.11.2 Resistance of the coil(R_m)

This is ohmic resistance of the moving coil. It is due to ρ , L and A. For an ammeter this should be as small as possible.

1.11.3 Sensitivity of the meter(S)

$$S = \frac{1}{I_{FSD}}(\Omega / volt), \uparrow S = \frac{Z \uparrow}{V}$$

It is also called ohms/volt rating of the instrument. Larger the sensitivity of an instrument, more accurate is the instrument. It is measured in Ω/volt. When the sensitivity is high, the impedance of meter is high. Hence it draws less current and loading affect is negligible. It is also defend as one over full scale deflection current.

1.12 Error in M.I instrument

1.12.1 Temperature error

Due to temperature variation, the resistance of the coil varies. This affects the deflection of the instrument. The coil should be made of manganin, so that the resistance is almost constant.

1.12.2 Hysteresis error

Due to hysteresis affect the reading of the instrument will not be correct. When the current is decreasing, the flux produced will not decrease suddenly. Due to this the meter reads a higher value of current. Similarly when the current increases the meter reads a lower value of current. This produces error in deflection. This error can be eliminated using small iron parts with narrow hysteresis loop so that the demagnetization takes place very quickly.

1.12.3 Eddy current error

The eddy currents induced in the moving iron affect the deflection. This error can be reduced by increasing the resistance of the iron.

1.12.4 Stray field error

Since the operating field is weak, the effect of stray field is more. Due to this, error is produced in deflection. This can be eliminated by shielding the parts of the instrument.

1.12.5 Frequency error

When the frequency changes the reactance of the coil changes.

$$Z = \sqrt{(R_m + R_S)^2 + X_L^2} \tag{1.70}$$

$$I = \frac{V}{Z} = \frac{V}{\sqrt{(R_m + R_S)^2 + X_L^2}} \tag{1.71}$$

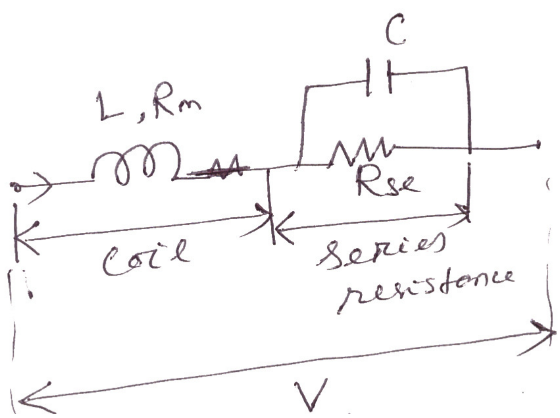


Fig. 1.18

Deflection of moving iron voltmeter depends upon the current through the coil. Therefore, deflection for a given voltage will be less at higher frequency than at low frequency. A capacitor is connected in parallel with multiplier resistance. The net reactance, ($X_L - X_C$) is very small, when compared to the series resistance. Thus the circuit impedance is made independent of frequency. This is because of the circuit is almost resistive.

$$C = 0.41 \frac{L}{(R_S)^2} \tag{1.72}$$

1.13 Electrostatic instrument

In multi cellular construction several vans and quadrants are provided. The voltage is to be measured is applied between the vanes and quadrant. The force of attraction between the vanes and quadrant produces a deflecting torque. Controlling torque is produced by spring control. Air friction damping is used.

The instrument is generally used for measuring medium and high voltage. The voltage is reduced to low value by using capacitor potential divider. The force of attraction is proportional to the square of the voltage.

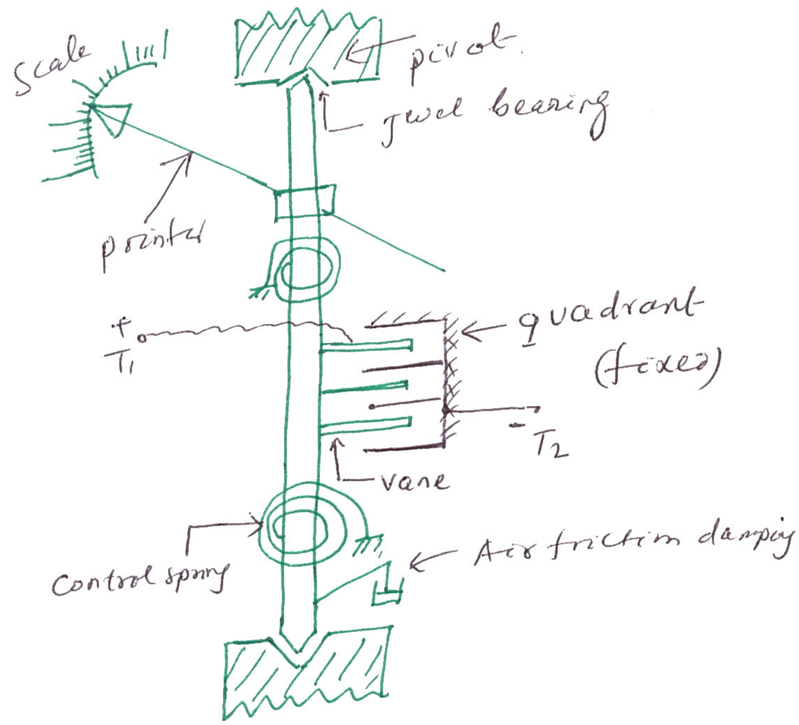


Fig. 1.19

<u>Torque develop by electrostatic instrument</u>

V=Voltage applied between vane and quadrant

C=capacitance between vane and quadrant

Energy stored=$\frac{1}{2}CV^2$ (1.73)

Let 'θ' be the deflection corresponding to a voltage V.

Let the voltage increases by dv, the corresponding deflection is'$\theta + d\theta$'

When the voltage is being increased, a capacitive current flows

$$i = \frac{dq}{dt} = \frac{d(CV)}{dt} = \frac{dC}{dt}V + C\frac{dV}{dt} \tag{1.74}$$

$V \times dt$ multiply on both side of equation (1.74)

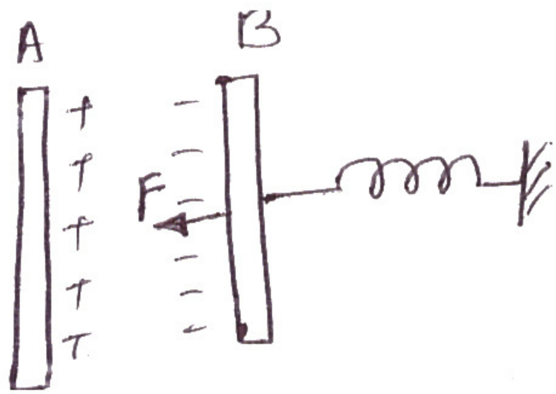


Fig. 1.20

$$Vidt = \frac{dC}{dt}V^2 dt + CV\frac{dV}{dt}dt \tag{1.75}$$

$$Vidt = V^2 dC + CVdV \tag{1.76}$$

$$\text{Change in stored energy} = \frac{1}{2}(C+dC)(V+dV)^2 - \frac{1}{2}CV^2 \tag{1.77}$$

$$= \frac{1}{2}\left[(C+dC)V^2 + dV^2 + 2VdV\right] - \frac{1}{2}CV^2$$

$$= \frac{1}{2}\left[CV^2 + CdV^2 + 2CVdV + V^2dC + dCdV^2 + 2VdVdC\right] - \frac{1}{2}CV^2$$

$$= \frac{1}{2}V^2 dC + CVdV$$

$$V^2 dC + CVdV = \frac{1}{2}V^2 dC + CVdV + F \times rd\theta \tag{1.78}$$

$$T_d \times d\theta = \frac{1}{2}V^2 dC \tag{1.79}$$

$$T_d = \frac{1}{2}V^2\left(\frac{dC}{d\theta}\right) \tag{1.80}$$

At steady state condition, $T_d = T_C$

$$K\theta = \frac{1}{2}V^2\left(\frac{dC}{d\theta}\right) \tag{1.81}$$

$$\theta = \frac{1}{2K}V^2\left(\frac{dC}{d\theta}\right) \tag{1.82}$$

Advantages

- ✓ It is used in both AC and DC.
- ✓ There is no frequency error.
- ✓ There is no hysteresis error.
- ✓ There is no stray magnetic field error. Because the instrument works on electrostatic principle.
- ✓ It is used for high voltage
- ✓ Power consumption is negligible.

Disadvantages

- ✓ Scale is not uniform
- ✓ Large in size
- ✓ Cost is more

1.14 Multi range Ammeter

When the switch is connected to position (1), the supplied current I_1

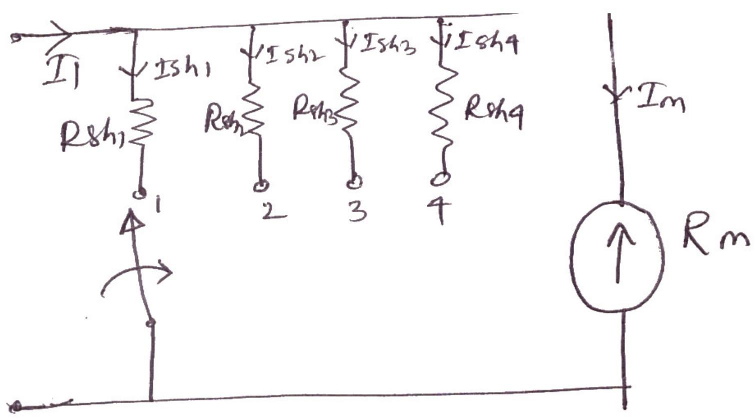


Fig. 1.21

$$I_{sh1}R_{sh1} = I_m R_m \tag{1.83}$$

$$R_{sh1} = \frac{I_m R_m}{I_{sh1}} = \frac{I_m R_m}{I_1 - I_m} \tag{1.84}$$

$$R_{sh1} = \frac{R_m}{\frac{I_1}{I_m} - 1}, R_{sh1} = \frac{R_m}{m_1 - 1}, m_1 = \frac{I_1}{I_m} = \text{Multiplying power of shunt}$$

$$R_{sh2} = \frac{R_m}{m_2 - 1}, m_2 = \frac{I_2}{I_m} \quad (1.85)$$

$$R_{sh3} = \frac{R_m}{m_3 - 1}, m_3 = \frac{I_3}{I_m} \quad (1.86)$$

$$R_{sh4} = \frac{R_m}{m_4 - 1}, m_4 = \frac{I_4}{I_m} \quad (1.87)$$

1.15 Ayrton shunt

$$R_1 = R_{sh1} - R_{sh2} \quad (1.88)$$

$$R_2 = R_{sh2} - R_{sh3} \quad (1.89)$$

$$R_3 = R_{sh3} - R_{sh4} \quad (1.90)$$

$$R_4 = R_{sh4} \quad (1.91)$$

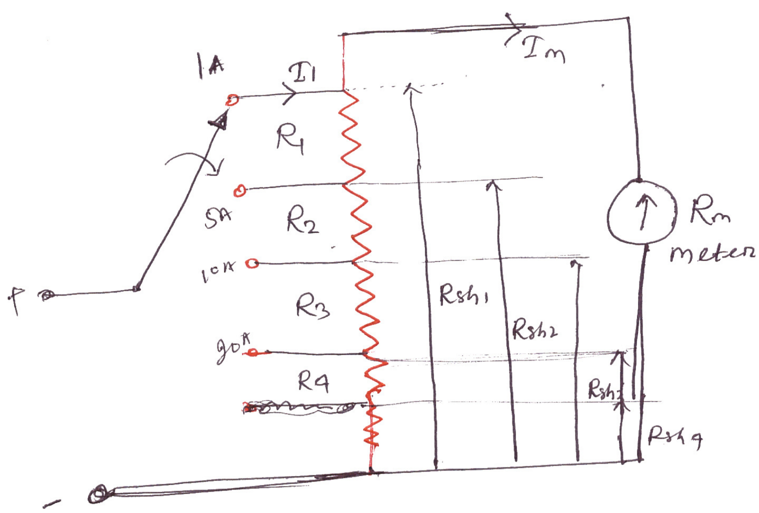


Fig. 1.22

Ayrton shunt is also called universal shunt. Ayrton shunt has more sections of resistance. Taps are brought out from various points of the resistor. The variable points in the o/p can be connected to any position. Various meters require different types of shunts. The Aryton shunt is used in the lab, so that any value of resistance between minimum and maximum specified can be used. It eliminates the possibility of having the meter in the circuit without a shunt.

1.16 Multi range D.C. voltmeter

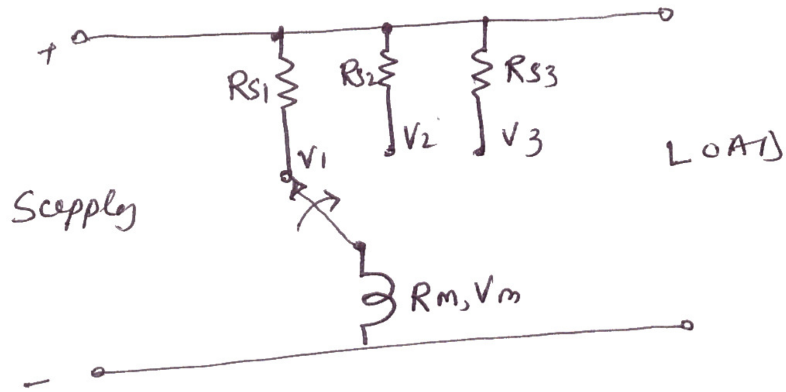


Fig. 1.23

$$R_{s1} = R_m(m_1 - 1)$$
$$R_{s2} = R_m(m_2 - 1) \qquad (1.92)$$
$$R_{s3} = R_m(m_3 - 1)$$

$$m_1 = \frac{V_1}{V_m}, m_2 = \frac{V_2}{V_m}, m_3 = \frac{V_3}{V_m} \qquad (1.93)$$

We can obtain different Voltage ranges by connecting different value of multiplier resistor in series with the meter. The number of these resistors is equal to the number of ranges required**.**

1.17 Potential divider arrangement

The resistance R_1, R_2, R_3 and R_4 is connected in series to obtained the ranges V_1, V_2, V_3 and V_4

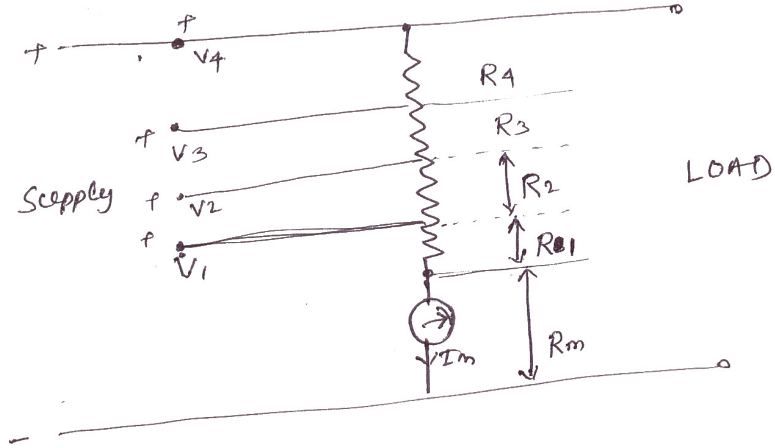


Fig. 1.24

Consider for voltage V_1, $(R_1 + R_m)I_m = V_1$

$$\therefore R_1 = \frac{V_1}{I_m} - R_m = \frac{V_1}{(\frac{V_m}{R_m})} - R_m = \left(\frac{V_1}{V_m}\right)R_m - R_m \tag{1.94}$$

$$R_1 = (m_1 - 1)R_m \tag{1.95}$$

For V_2, $(R_2 + R_1 + R_m)I_m = V_2 \Rightarrow R_2 = \frac{V_2}{I_m} - R_1 - R_m$ (1.96)

$$R_2 = \frac{V_2}{\left(\frac{V_m}{R_m}\right)} - (m_1 - 1)R_m - R_m \tag{1.97}$$

$$R_2 = m_2 R_m - R_m - (m_1 - 1)R_m$$

$$= R_m(m_2 - 1 - m_1 + 1) \tag{1.98}$$

$$R_2 = (m_2 - m_1)R_m \tag{1.99}$$

For V_3 $(R_3 + R_2 + R_1 + R_m)I_m = V_3$

$$R_3 = \frac{V_3}{I_m} - R_2 - R_1 - R_m$$

$$= \frac{V_3}{V_m}R_m - (m_2 - m_1)R_m - (m_1 - 1)R_m - R_m$$

$$= m_3 R_m - (m_2 - m_1)R_m - (m_1 - 1)R_m - R_m$$

$$R_3 = (m_3 - m_2)R_m$$

For V_4 $\quad (R_4 + R_3 + R_2 + R_1 + R_m)I_m = V_4$

$$R_4 = \frac{V_4}{I_m} - R_3 - R_2 - R_1 - R_m$$

$$= \left(\frac{V_4}{V_m}\right)R_m - (m_3 - m_2)R_m - (m_2 - m_1)R_m - (m_1 - 1)R_m - R_m$$

$$R_4 = R_m[m_4 - m_3 + m_2 - m_2 + m_1 - m_1 + 1 - 1]$$
$$R_4 = (m_4 - m_3)R_m$$

Example: 1.1

A PMMC ammeter has the following specification

Coil dimension are 1cm$\times$ 1cm. Spring constant is $0.15\times10^{-6} N-m/rad$, Flux density is $1.5\times10^{-3} wb/m^2$.Determine the no. of turns required to produce a deflection of 90^0 when a current 2mA flows through the coil.

Solution:

At steady state **condition** $T_d = T_C$

$BANI = K\theta$

$\Rightarrow N = \frac{K\theta}{BAI}$

A=$1\times10^{-4} m^2$

K=$0.15\times10^{-6} \frac{N-m}{rad}$

B=$1.5\times10^{-3} wb/m^2$

I=$2\times10^{-3} A$

$\theta = 90^\circ = \frac{\Pi}{2} rad$

N=785 ans.

Example: 1.2

The pointer of a moving coil instrument gives full scale deflection of 20mA. The potential difference across the meter when carrying 20mA is 400mV.The instrument to be used is 200A for full scale deflection. Find the shunt resistance required to achieve this, if the instrument to be used as a voltmeter for full scale reading with 1000V. Find the series resistance to be connected it?

Solution:

Case-1

V_m =400mV

$I_m = 20mA$

I=200A

$$R_m = \frac{V_m}{I_m} = \frac{400}{20} = 20\Omega$$

$$I = I_m\left(1 + \frac{R_m}{R_{sh}}\right)$$

$$200 = 20 \times 10^{-3}\left[1 + \frac{20}{R_{sh}}\right]$$

$$R_{sh} = 2 \times 10^{-3}\Omega$$

Case-II

V=1000V

$$V = V_m\left(1 + \frac{R_{se}}{R_m}\right)$$

$$4000 = 400 \times 10^{-3}\left(1 + \frac{R_{se}}{20}\right)$$

$$R_{se} = 49.98k\Omega$$

Example: 1.3

A 150 v moving iron voltmeter is intended for 50HZ, has a resistance of 3kΩ. Find the series resistance required to extent the range of instrument to 300v. If the 300V instrument is used to measure a d.c. voltage of 200V. Find the voltage across the meter?

Solution:

$R_m = 3k\Omega, V_m = 150V, V = 300V$

$$V = V_m\left(1+\frac{R_{se}}{R_m}\right)$$

$$300 = 150\left(1+\frac{R_{se}}{3}\right) \Rightarrow R_{se} = 3k\Omega$$

Case-II $\quad V = V_m\left(1+\frac{R_{se}}{R_m}\right)$

$$200 = V_m\left(1+\frac{3}{3}\right)$$

$\therefore V_m = 100V$ Ans

Example: 1.4

What is the value of series resistance to be used to extent '0'to 200V range of 20,000Ω/volt voltmeter to 0 to 2000 volt?

Solution:

$$V_{se} = V - V = 1800$$

$$I_{FSD} = \frac{1}{20000} = \frac{1}{Sensitivity}$$

$V_{se} = R_{se} \times i_{FSD} \Rightarrow R_{se} = 36M\Omega$ ans.

Example: 1.5

A moving coil instrument whose resistance is 25Ω gives a full scale deflection with a current of 1mA. This instrument is to be used with a manganin shunt, to extent its range to 100mA. Calculate the error caused by a 10^0C rise in temperature when:

(a) Copper moving coil is connected directly across the manganin shunt.

(b) A 75 ohm manganin resistance is used in series with the instrument moving coil.

The temperature co-efficient of copper is $0.004/^0$C and that of manganin is 0.00015^0/C.

Solution:

Case-1

$$I_m = 1mA$$

$$R_m = 25\Omega$$

I=100mA

$$I = I_m\left(1+\frac{R_m}{R_{sh}}\right)$$

$$100 = 1\left(1+\frac{25}{R_{sh}}\right) \Rightarrow \frac{25}{R_{sh}} = 99$$

$$\Rightarrow R_{sh} = \frac{25}{99} = 0.2525\Omega$$

Instrument resistance for 10^0C rise in temperature, $R_{mt} = 25(1+0.004\times10)$

$$R_t = R_o(1+\rho_t \times t)$$

$$R_{m/t=10^\circ} = 26\Omega$$

Shunt resistance for 10^0C, rise in temperature

$$R_{sh/t=10^\circ} = 0.2525(1+0.00015\times10) = 0.2529\Omega$$

Current through the meter for 100mA in the main circuit for 10^0C rise in temperature

$$I = I_m\left(1+\frac{R_m}{R_{sh}}\right)\Bigg|_{t=10^\circ C}$$

$$100 = I_{mt}\left(1+\frac{26}{0.2529}\right)$$

$$I_m\big|_{t=10} = 0.963mA$$

But normal meter current=1mA

Error due to rise in temperature=(0.963-1)*100=-3.7%

Case-b As voltmeter

Total resistance in the meter circuit= $R_m + R_{sh} = 25+75 = 100\Omega$

$$I = I_m\left(1+\frac{R_m}{R_{sh}}\right)$$

$$100 = 1\left(1+\frac{100}{R_{sh}}\right)$$

$$R_{sh} = \frac{100}{100-1} = 1.01\Omega$$

Resistance of the instrument circuit for 10^0C rise in temperature

$$R_m\big|_{t=10} = 25(1+0.004\times10)+75(1+0.00015\times10)=101.11\Omega$$

Shunt resistance for 10^0C rise in temperature

$$R_{sh}\big|_{t=10} = 1.01(1+0.00015\times10)=1.0115\Omega$$

$$I = I_m\left(1+\frac{R_m}{R_{sh}}\right)$$

$$100 = I_m\left(1+\frac{101.11}{1.0115}\right)$$

$$I_m\Big|t=10^\circ = 0.9905mA$$

Error =(0.9905-1)*100=-0.95%

Example: 1.6

The coil of a 600V M.I meter has an inductance of 1 henery. It gives correct reading at 50HZ and requires 100mA. For its full scale deflection, what is % error in the meter when connected to 200V D.C. by comparing with 200V A.C?

Solution:

$$V_m = 600V, I_m = 100mA$$

Case-I A.C.

$$Z_m = \frac{V_m}{I_m} = \frac{600}{0.1} = 6000\Omega$$

$$X_L = 2\Pi fL = 314\Omega$$

$$R_m = \sqrt{{Z_m}^2 - X_L^2} = \sqrt{(6000)^2-(314)^2} = 5990\Omega$$

$$I_{AC} = \frac{V_{AC}}{Z} = \frac{200}{6000} = 33.33mA$$

Case-II D.C

$$I_{DC} = \frac{V_{DC}}{R_m} = \frac{200}{5990} = 33.39mA$$

$$\text{Error}=\frac{I_{DC}-I_{AC}}{I_{AC}}\times 100=\frac{33.39-33.33}{33.33}\times 100=0.18\%$$

Example: 1.7

A 250V M.I. voltmeter has coil resistance of 500Ω, coil inductance 0f 1.04 H and series resistance of 2kΩ. The meter reads correctively at 250V D.C. What will be the value of capacitance to be used for shunting the series resistance to make the meter read correctly at 50HZ? What is the reading of voltmeter on A.C. without capacitance?

Solution: $C=0.41\dfrac{L}{(R_S)^2}$

$$=0.41\times\frac{1.04}{(2\times 10^3)^2}=0.1\mu F$$

For A.C $Z=\sqrt{(R_m+R_{Se})^2+X_L^2}$

$$Z=\sqrt{(500+2000)^2+(314)^2}=2520\Omega$$

With D.C

$R_{total}=2500\Omega$

For $2500\Omega\rightarrow 250\text{V}$

$$1\Omega\rightarrow\frac{250}{2500}$$

$$2520\Omega\rightarrow\frac{250}{2500}\times 2520=248V$$

Example: 1.8

The relationship between inductance of moving iron ammeter, the current and the position of pointer is as follows:

Reading (A)	1.2	1.4	1.6	1.8
Deflection (degree)	36.5	49.5	61.5	74.5
Inductance (μH)	575.2	576.5	577.8	578.8

Calculate the deflecting torque and the spring constant when the current is 1.5A?

Solution:

For current I=1.5A, θ=55.5 degree=0.96865 rad

$$\frac{dL}{d\theta}=\frac{577.65-576.5}{60-49.5}=0.11\mu H/\deg ree=6.3\mu H/rad$$

Deflecting torque , $T_d=\frac{1}{2}I^2\frac{dL}{d\theta}=\frac{1}{2}(1.5)^2\times 6.3\times 10^{-6}=7.09\times 10^{-6}\,N-m$

Spring constant, $K=\frac{T_d}{\theta}=\frac{7.09\times 10^{-6}}{0.968}=7.319\times 10^{-6}\,\frac{N-m}{rad}$

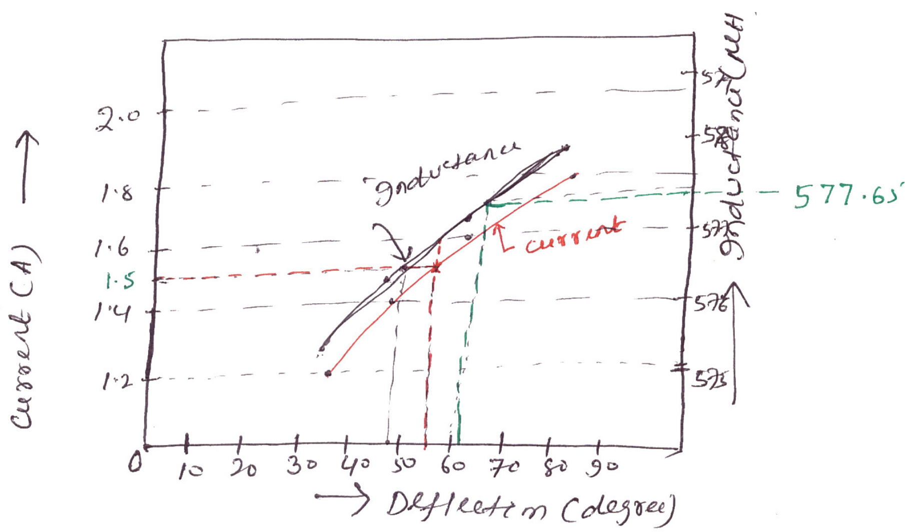


Fig. 1.25

Example: 1.9

For a certain dynamometer ammeter the mutual inductance 'M' varies with deflection θ as $M=-6\cos(\theta+30^{\circ})mH$ .Find the deflecting torque produced by a direct current of 50mA corresponding to a deflection of 60^0.

Solution:

$$T_d=I_1I_2\frac{dM}{d\theta}=I^2\frac{dM}{d\theta}$$

$$M=-6\cos(\theta+30^{\circ})$$

$$\frac{dM}{d\theta}=6\sin(\theta+30)mH$$

$$\frac{dM}{d\theta}\Big|_{\theta=60}=6\sin 90=6mH/\deg$$

$$T_d=I^2\frac{dM}{d\theta}=(50\times 10^{-3})^2\times 6\times 10^{-3}=15\times 10^{-6}\,N-m$$

Example: 1.10

The inductance of a moving iron ammeter with a full scale deflection of 90^0 at 1.5A, is given by the expression $L = 200 + 40\theta - 4\theta^2 - \theta^3 \mu H$, where θ is deflection in radian from the zero position. Estimate the angular deflection of the pointer for a current of 1.0A.

Solution:

$$L = 200 + 40\theta - 4\theta^2 - \theta^3 \mu H$$

$$\left.\frac{dL}{d\theta}\right|_{\theta=90^\circ} = 40 - 8\theta - 3\theta^2 \mu H / rad$$

$$\left.\frac{dL}{d\theta}\right|_{\theta=90^\circ} = 40 - 8\times\frac{\Pi}{2} - 3(\frac{\Pi}{2})^2 \mu H / rad = 20\mu H / rad$$

$$\therefore \theta = \frac{1}{2K} I^2 \left(\frac{dL}{d\theta}\right)$$

$$\frac{\Pi}{2} = \frac{1}{2}\frac{(1.5)^2}{K} \times 20\times10^{-6}$$

K=Spring constant=$14.32\times10^{-6} N - m / rad$

For I=1A, $\therefore \theta = \frac{1}{2K} I^2 \left(\frac{dL}{d\theta}\right)$

$$\therefore \theta = \frac{1}{2}\times\frac{(1)^2}{14.32\times10^{-6}}\left(40 - 8\theta - 3\theta^2\right)$$

$$3\theta + 36.64\theta^2 - 40 = 0$$

$$\theta = 1.008 rad, 57.8^\circ$$

Example: 1.11

The inductance of a moving iron instrument is given by $L = 10 + 5\theta - \theta^2 - \theta^3 \mu H$, where θ is the deflection in radian from zero position. The spring constant is $12\times10^{-6} N - m / rad$. Estimate the deflection for a current of 5A.

Solution:

$$\frac{dL}{d\theta} = (5-2\theta)\frac{\mu H}{rad}$$

$$\therefore \theta = \frac{1}{2K} I^2 \left(\frac{dL}{d\theta}\right)$$

$$\therefore \theta = \frac{1}{2} \times \frac{(5)^2}{12\times 10^{-6}}(5-2\theta)\times 10^{-6}$$

$$\therefore \theta = 1.69 rad, 96.8^{\circ}$$

Example: 1.12

The following figure gives the relation between deflection and inductance of a moving iron instrument.

Deflection (degree)	20	30	40	50	60	70	80	90
Inductance (μH)	335	345	355.5	366.5	376.5	385	391.2	396.5

Find the current and the torque to give a deflection of (a) 30^0 (b) 80^0 . Given that control spring constant is $0.4\times 10^{-6} N-m/\deg ree$

Solution:

$$\theta = \frac{1}{2K} I^2 \left(\frac{dL}{d\theta}\right)$$

(a) For $\theta = 30^{\circ}$

The curve is linear

$$\therefore \left(\frac{dL}{d\theta}\right)_{\theta=30} = \frac{355.5-335}{40-20} = 1.075\mu H/\deg ree = 58.7\mu H/rad$$

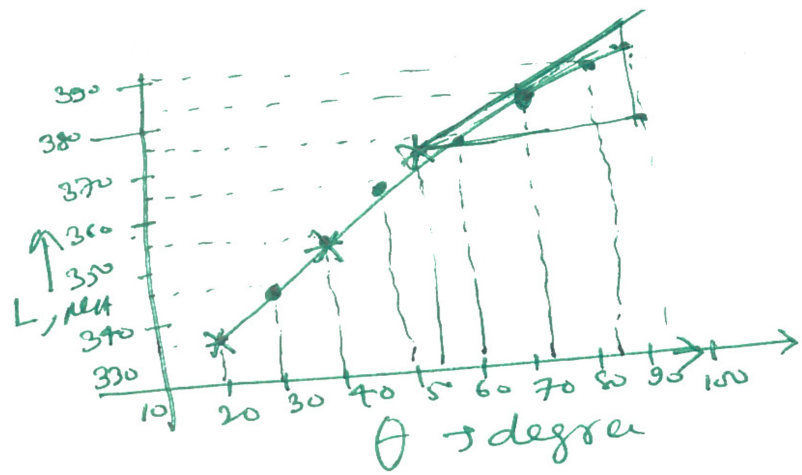


Fig. 1.26

Example: 1.13

In an electrostatic voltmeter the full scale deflection is obtained when the moving plate turns through 90^0. The torsional constant is $10\times10^{-6}\,N-m/rad$. The relation between the angle of deflection and capacitance between the fixed and moving plates is given by

Deflection (degree)	0	10	20	30	40	50	60	70	80	90
Capacitance (PF)	81.4	121	156	189.2	220	246	272	294	316	334

Find the voltage applied to the instrument when the deflection is 90^0?

Solution:

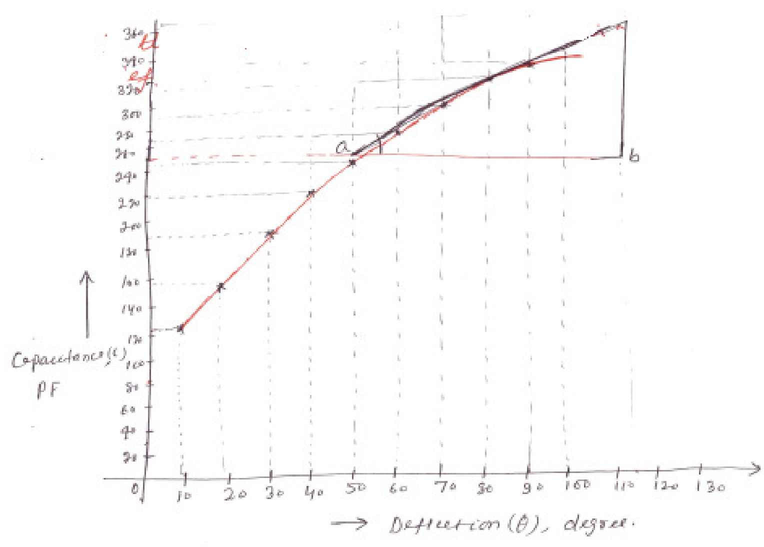


Fig. 1.27

$$\frac{dC}{d\theta} = \tan\theta = \frac{bc}{ab} = \frac{370-250}{110-44} = 1.82PF/\deg ree = 104.2PF/rad$$

Spring constant $K = 10\times10^{-6}\frac{N-m}{rad} = 0.1745\times10^{-6}N-m/\deg ree$

$$\theta = \frac{1}{2K}V^2\left(\frac{dC}{d\theta}\right) \Rightarrow V = \sqrt{\frac{2K\theta}{\frac{dC}{d\theta}}}$$

$$V = \sqrt{\frac{2\times0.1745\times10^{-6}\times90}{104.2\times10^{-12}}} = 549volt$$

Example: 1.14

Design a multi range d.c. mille ammeter using a basic movement with an internal resistance $R_m = 50\Omega$ and a full scale deflection current $I_m = 1mA$. The ranges required are 0-10mA; 0-50mA; 0-100mA and 0-500mA.

Solution:

Case-I 0-10mA

Multiplying power $m = \frac{I}{I_m} = \frac{10}{1} = 10$

$\therefore$ Shunt resistance $R_{sh1} = \frac{R_m}{m-1} = \frac{50}{10-1} = 5.55\Omega$

Case-II 0-50mA

$$m = \frac{50}{1} = 50$$

$$R_{sh2} = \frac{R_m}{m-1} = \frac{50}{50-1} = 1.03\Omega$$

Case-III 0-100mA, $m = \frac{100}{1} = 100\Omega$

$$R_{sh3} = \frac{R_m}{m-1} = \frac{50}{100-1} = 0.506\Omega$$

Case-IV 0-500mA, $m = \frac{500}{1} = 500\Omega$

$$R_{sh4} = \frac{R_m}{m-1} = \frac{50}{500-1} = 0.1\Omega$$

Example: 1.15

A moving coil voltmeter with a resistance of 20Ω gives a full scale deflection of 120^0, when a potential difference of 100mV is applied across it. The moving coil has dimension of 30mm*25mm and is wounded with 100 turns. The control spring constant is $0.375\times10^{-6} N-m/\deg ree.$ Find the flux density, in the air gap. Find also the diameter of copper wire of coil winding if 30% of instrument resistance is due to coil winding. The specific resistance for copper=$1.7\times10^{-8}\Omega m$.

Solution:

Data given

$V_m = 100mV$

$R_m = 20\Omega$

$\theta = 120^{\circ}$

N=100

$K = 0.375\times10^{-6} N-m/\deg ree$

$R_C = 30\% of R_m$

$\rho = 1.7\times10^{-8}\Omega m$

$$I_m = \frac{V_m}{R_m} = 5\times10^{-3} A$$

$$T_d = BANI, T_C = K\theta = 0.375\times10^{-6}\times120 = 45\times10^{-6} N-m$$

$$B = \frac{T_d}{ANI} = \frac{45\times10^{-6}}{30\times25\times10^{-6}\times100\times5\times10^{-3}} = 0.12wb/m^2$$

$R_C = 0.3\times20 = 6\Omega$

Length of mean turn path =2(a+b) =2(55)=110mm

$$R_C = N\left(\frac{\rho l}{A}\right)$$

$$A = \frac{N\times\rho\times(l_t)}{R_C} = \frac{100\times1.7\times10^{-8}\times110\times10^{-3}}{6}$$

$= 3.116 \times 10^{-8} m^2$

$= 31.16 \times 10^{-3} mm^2$

$A = \frac{\Pi}{4} d^2 \Rightarrow d = 0.2mm$

Example: 1.16

A moving coil instrument gives a full scale deflection of 10mA, when the potential difference across its terminal is 100mV. Calculate

(1) The shunt resistance for a full scale deflection corresponding to 100A

(2) The resistance for full scale reading with 1000V.

Calculate the power dissipation in each case?

Solution:

Data given

$I_m = 10mA$

$V_m = 100mV$

$I = 100A$

$$I = I_m \left(1 + \frac{R_m}{R_{sh}} \right)$$

$$100 = 10 \times 10^{-3} \left(1 + \frac{10}{R_{sh}} \right)$$

$R_{sh} = 1.001 \times 10^{-3} \Omega$

$R_{se} = ??, V = 1000V$

$$R_m = \frac{V_m}{I_m} = \frac{100}{10} = 10\Omega$$

$$V = V_m \left(1 + \frac{R_{se}}{R_m} \right)$$

$$1000 = 100 \times 10^{-3} \left(1 + \frac{R_{se}}{10} \right)$$

$\therefore R_{se} = 99.99 K\Omega$

Example: 1.17

Design an Aryton shunt to provide an ammeter with current ranges of 1A,5A,10A and 20A. A basic meter with an internal resistance of 50w and a full scale deflection current of 1mA is to be used.

Solution: Data given

$$
\begin{array}{l|l|l}
 & & m_1 = \dfrac{I_1}{I_m} = 1000A \\
I_m = 1\times10^{-3}A & I_1 = 1A & m_2 = \dfrac{I_2}{I_m} = 5000A \\
R_m = 50\Omega & I_2 = 5A & m_3 = \dfrac{I_3}{I_m} = 10000A \\
 & I_3 = 10A & m_4 = \dfrac{I_4}{I_m} = 20000A \\
 & I_4 = 20A &
\end{array}
$$

$$R_{sh1} = \frac{R_m}{m_1 - 1} = \frac{50}{1000-1} = 0.05\Omega$$

$$R_{sh2} = \frac{R_m}{m_2 - 1} = \frac{50}{5000-1} = 0.01\Omega$$

$$R_{sh3} = \frac{R_m}{m_3 - 1} = \frac{50}{10000-1} = 0.005\Omega$$

$$R_{sh4} = \frac{R_m}{m_4 - 1} = \frac{50}{20000-1} = 0.0025\Omega$$

$\therefore$ The resistances of the various section of the universal shunt are

$$R_1 = R_{sh1} - R_{sh2} = 0.05 - 0.01 = 0.04\Omega$$

$$R_2 = R_{sh2} - R_{sh3} = 0.01 - 0.005 = 0.005\Omega$$

$$R_3 = R_{sh3} - R_{sh4} = 0.005 - 0.025 = 0.0025\Omega$$

$$R_4 = R_{sh4} = 0.0025\Omega$$

Example: 1.18

A basic d' Arsonval meter movement with an internal resistance $R_m = 100\Omega$ and a full scale current of $I_m = 1mA$ is to be converted in to a multi range d.c. voltmeter with ranges of 0-10V, 0-50V, 0-250V, 0-500V. Find the values of various resistances using the potential divider arrangement.

Solution:

Data given

$$R_m = 100\Omega$$
$$I_m = 1mA$$
$$V_m = I_m \times R_m$$
$$V_m = 100 \times 1 \times 10^{-3}$$
$$V_m = 100mV$$

$$m_1 = \frac{V_1}{V_m} = \frac{10}{100 \times 10^{-3}} = 100$$
$$m_2 = \frac{V_2}{V_m} = \frac{50}{100 \times 10^{-3}} = 500$$
$$m_3 = \frac{V_3}{V_m} = \frac{250}{100 \times 10^{-3}} = 2500$$
$$m_4 = \frac{V_4}{V_m} = \frac{500}{100 \times 10^{-3}} = 5000$$

$$R_1 = (m_1 - 1)R_m = (100 - 1) \times 100 = 9900\Omega$$
$$R_2 = (m_2 - m_1)R_m = (500 - 100) \times 100 = 40K\Omega$$
$$R_3 = (m_3 - m_2)R_m = (2500 - 500) \times 100 = 200K\Omega$$
$$R_4 = (m_4 - m_3)R_m = (5000 - 2500) \times 100 = 250K\Omega$$

AC BRIDGES

2.1 General form of A.C. bridge

AC bridge are similar to D.C. bridge in topology(way of connecting).It consists of four arm AB,BC,CD and DA .Generally the impedance to be measured is connected between 'A' and 'B'. A detector is connected between 'B' and 'D'. The detector is used as null deflection instrument. Some of the arms are variable element. By varying these elements, the potential values at 'B' and 'D' can be made equal. This is called balancing of the bridge.

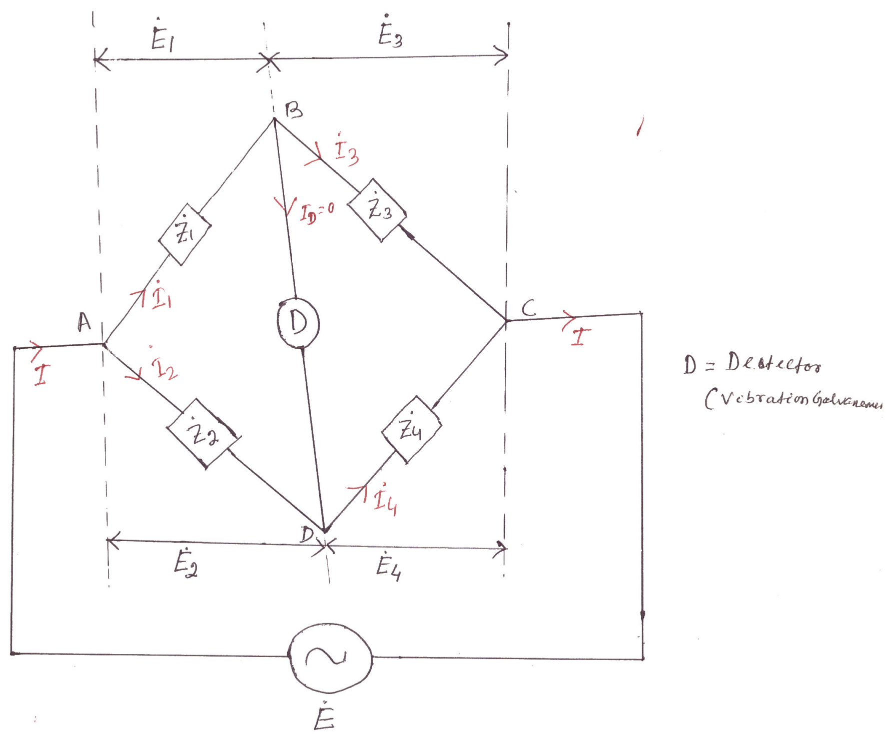


Fig. 2.1 General form of A.C. bridge

At the balance condition, the current through detector is zero.

$$\therefore \dot{I}_1 = \dot{I}_3$$

$$\dot{I}_2 = \dot{I}_4$$

$$\therefore \frac{\dot{I}_1}{\dot{I}_2} = \frac{\dot{I}_3}{\dot{I}_4} \tag{2.1}$$

At balance condition,

Voltage drop across ‘AB’=voltage drop across ‘AD’.

$$\dot{E}_1 = \dot{E}_2$$

$$\therefore \dot{I}_1 \dot{Z}_1 = \dot{I}_2 \dot{Z}_2 \qquad (2.2)$$

Similarly, Voltage drop across ‘BC’=voltage drop across ‘DC’

$$\dot{E}_3 = \dot{E}_4$$

$$\therefore \dot{I}_3 \dot{Z}_3 = \dot{I}_4 \dot{Z}_4 \qquad (2.3)$$

From Eqn. (2.2), we have $\therefore \frac{\dot{I}_1}{\dot{I}_2} = \frac{\dot{Z}_2}{\dot{Z}_1}$ (2.4)

From Eqn. (2.3), we have $\therefore \frac{\dot{I}_3}{\dot{I}_4} = \frac{\dot{Z}_4}{\dot{Z}_3}$ (2.5)

From equation -2.1, it can be seen that, equation -2.4 and equation-2.5 are equal.

$$\therefore \frac{\dot{Z}_2}{\dot{Z}_1} = \frac{\dot{Z}_4}{\dot{Z}_3}$$

$$\therefore \dot{Z}_1 \dot{Z}_4 = \dot{Z}_2 \dot{Z}_3$$

Products of impedances of opposite arms are equal.

$$\therefore |Z_1|\angle\theta_1 |Z_4|\angle\theta_4 = |Z_2|\angle\theta_2 |Z_3|\angle\theta_3$$

$$\Rightarrow |Z_1||Z_4|\angle\theta_1 + \theta_4 = |Z_2||Z_3|\angle\theta_2 + \theta_3$$

$$|Z_1||Z_4| = |Z_2||Z_3|$$

$$\theta_1 + \theta_4 = \theta_2 + \theta_3$$

* For balance condition, magnitude on either side must be equal.

* Angle on either side must be equal.

Summary

For balance condition,

- $\dot{I}_1 = \dot{I}_3, \quad \dot{I}_2 = \dot{I}_4$
- $|Z_1||Z_4| = |Z_2||Z_3|$
- $\theta_1 + \theta_4 = \theta_2 + \theta_3$
- $\dot{E}_1 = \dot{E}_2$ & $\dot{E}_3 = \dot{E}_4$

2.2 Types of detector

The following types of instruments are used as detector in A.C. bridge.

- Vibration galvanometer
- Head phones (speaker)
- Tuned amplifier

2.2.1 Vibration galvanometer

Between the point 'B' and 'D' a vibration galvanometer is connected to indicate the bridge balance condition. This A.C. galvanometer which works on the principle of resonance. The A.C. galvanometer shows a dot, if the bridge is unbalanced.

2.2.2 Head phones

Two speakers are connected in parallel in this system. If the bridge is unbalanced, the speaker produced more sound energy. If the bridge is balanced, the speaker do not produced any sound energy.

2.2.3 Tuned amplifier

If the bridge is unbalanced the output of tuned amplifier is high. If the bridge is balanced, output of amplifier is zero.

2.3 Measurements of inductance

2.3.1 Maxwell's inductance bridge

The choke for which R_1 and L_1 have to measure connected between the points 'A' and 'B'. In this method the unknown inductance is measured by comparing it with the standard inductance.

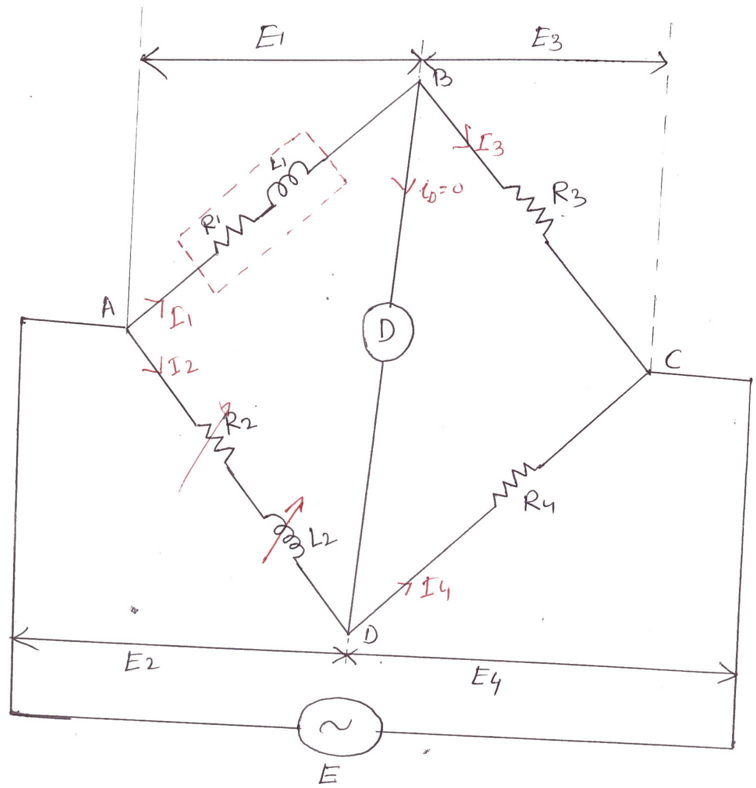


Fig. 2.2 Maxwell's inductance bridge

L_2 is adjusted, until the detector indicates zero current.

Let R_1= unknown resistance

L_1= unknown inductance of the choke.

L_2= known standard inductance

R_1,R_2,R_4= known resistances.

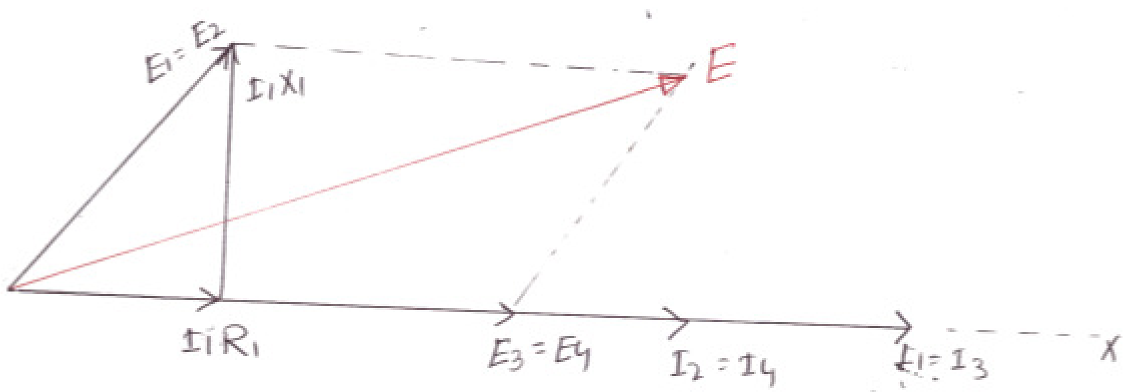


Fig 2.3 Phasor diagram of Maxwell's inductance bridge

At balance condition, $\dot{Z}_1 \dot{Z}_4 = \dot{Z}_2 \dot{Z}_3$

$$(R_1 + jXL_1)R_4 = (R_2 + jXL_2)R_3$$

$$(R_1 + jwL_1)R_4 = (R_2 + jwL_2)R_3$$

$$R_1R_4 + jwL_1R_4 = R_2R_3 + jwL_2R_3$$

Comparing real part,

$$R_1R_4 = R_2R_3$$

$$\therefore R_1 = \frac{R_2R_3}{R_4} \tag{2.6}$$

Comparing the imaginary parts,

$$wL_1R_4 = wL_2R_3$$

$$L_1 = \frac{L_2R_3}{R_4} \tag{2.7}$$

Q-factor of choke, $Q = \dfrac{WL_1}{R_1} = \dfrac{WL_2R_3R_4}{R_4R_2R_3}$

$$Q = \frac{WL_2}{R_2} \tag{2.8}$$

Advantages

- ✓ Expression for R_1 and L_1 are simple.
- ✓ Equations area simple
- ✓ They do not depend on the frequency (as w is cancelled)
- ✓ R_1 and L_1 are independent of each other.

Disadvantages

- ✓ Variable inductor is costly.
- ✓ Variable inductor is bulky.

2.3.2 Maxwell's inductance capacitance bridge

Unknown inductance is measured by comparing it with standard capacitance. In this bridge, balance condition is achieved by varying 'C_4'.

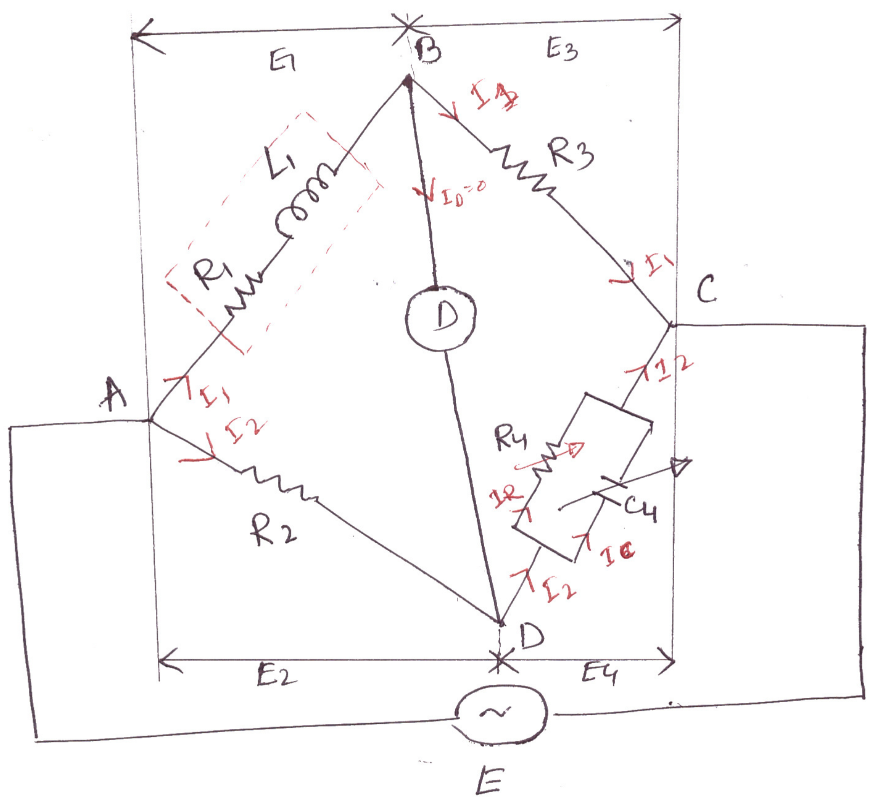


Fig 2.4 Maxwell's inductance capacitance bridge

At balance condition, $Z_1Z_4=Z_3Z_2$ (2.9)

$$Z_4 = R_4 \parallel \frac{1}{jwC_4} = \frac{R_4 \times \frac{1}{jwC_4}}{R_4 + \frac{1}{jwC_4}}$$

$$Z_4 = \frac{R_4}{jwR_4C_4 + 1} = \frac{R_4}{1 + jwR_4C_4} \quad (2.10)$$

$\therefore$ Substituting the value of Z_4 from eqn. (2.10) in eqn. (2.9) we get

$$(R_1 + jwL_1) \times \frac{R_4}{1 + jwR_4C_4} = R_2R_3$$

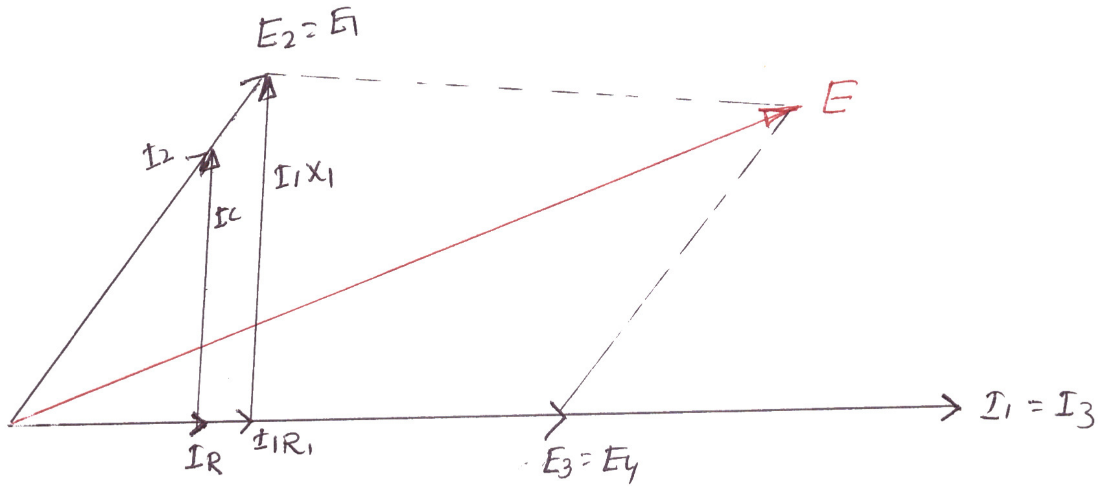


Fig 2.5 Phasor diagram of Maxwell's inductance capacitance bridge

$$(R_1 + jwL_1)R_4 = R_2R_3(1 + jwR_4C_4)$$

$$R_1R_4 + jwL_1R_4 = R_2R_3 + jwC_4R_4R_2R_3$$

Comparing real parts,

$$R_1R_4 = R_2R_3$$

$$\Rightarrow R_1 = \frac{R_2 R_3}{R_4} \tag{2.11}$$

Comparing imaginary part,

$$wL_1 R_4 = wC_4 R_4 R_2 R_3$$

$$L_1 = C_4 R_2 R_3 \tag{2.12}$$

Q-factor of choke,

$$Q = \frac{WL_1}{R_1} = w \times C_4 R_2 R_3 \times \frac{R_4}{R_2 R_3}$$

$$Q = wC_4 R_4 \tag{2.13}$$

Advantages

- ✓ Equation of L_1 and R_1 are simple.
- ✓ They are independent of frequency.
- ✓ They are independent of each other.
- ✓ Standard capacitor is much smaller in size than standard inductor.

Disadvantages

- ✓ Standard variable capacitance is costly.
- ✓ It can be used for measurements of Q-factor in the ranges of 1 to 10.
- ✓ It cannot be used for measurements of choke with Q-factors more than 10.

  We know that $Q = wC_4R_4$

For measuring chokes with higher value of Q-factor, the value of C_4 and R_4 should be higher. Higher values of standard resistance are very expensive. Therefore this bridge cannot be used for higher value of Q-factor measurements.

2.3.3 Hay's bridge

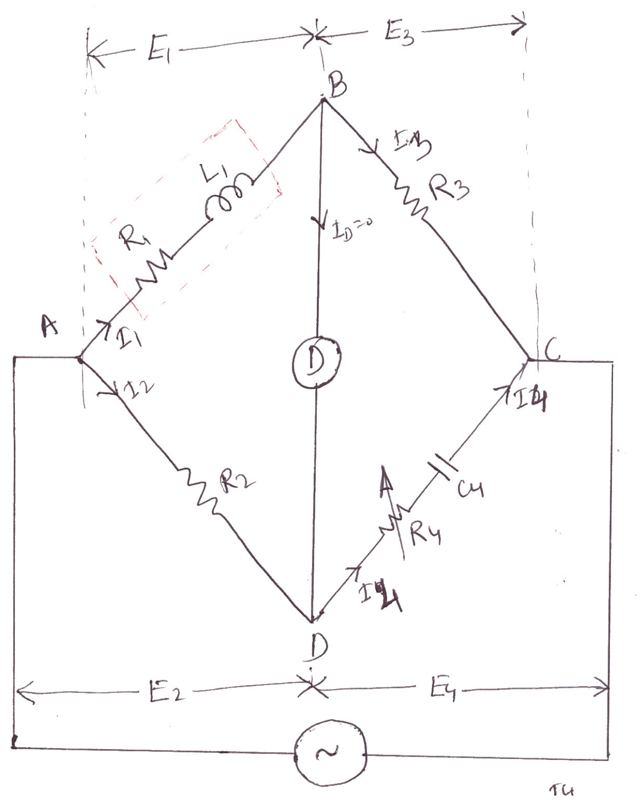


Fig 2.6 Hay's bridge

- $\dot{E}_1 = I_1 R_1 + jI_1 X_1$
- $\dot{E} = \dot{E}_1 + \dot{E}_3$
- $\dot{E}_4 = \dot{I}_4 R_4 + \dfrac{I_4}{jwC_4}$
- $\dot{E}_3 = I_3 R_3$

$$Z_4 = R_4 + \frac{1}{jwC_4} = \frac{1 + jwR_4C_4}{jwC_4}$$

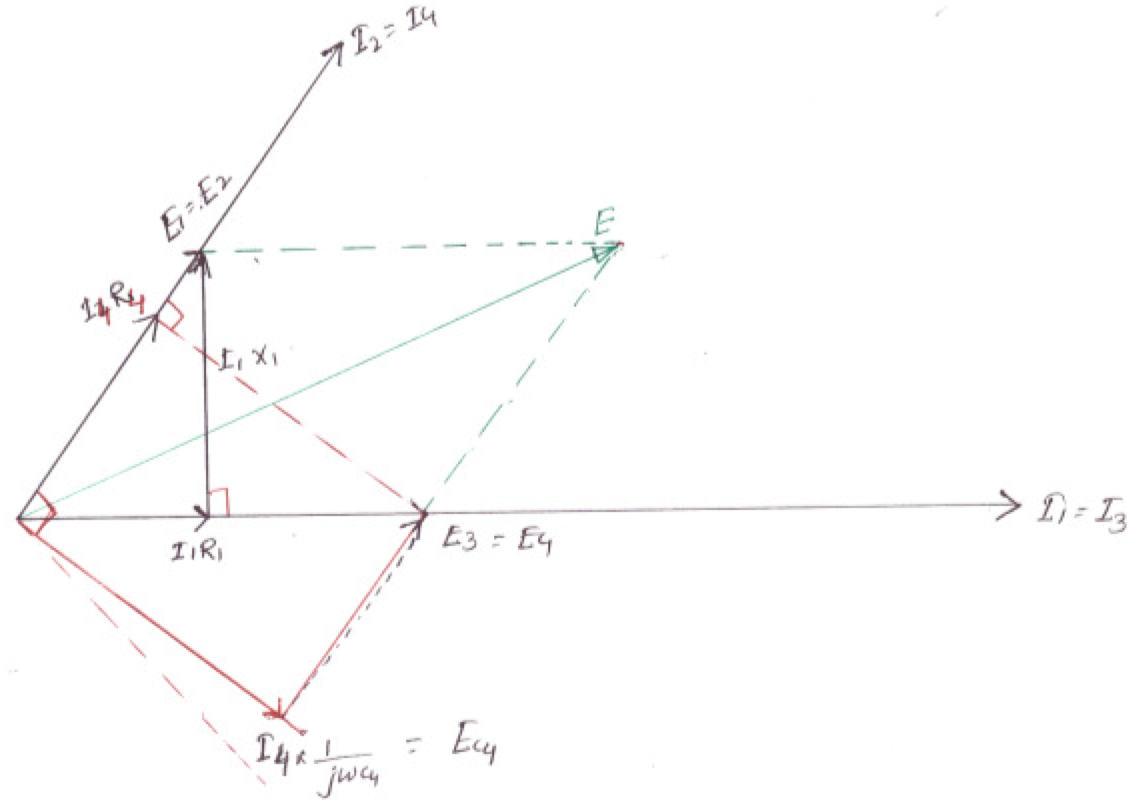


Fig 2.7 Phasor diagram of Hay's bridge

At balance condition, $Z_1Z_4=Z_3Z_2$

$$(R_1 + jwL_1)(\frac{1 + jwR_4C_4}{jwC_4}) = R_2R_3$$

$$(R_1 + jwL_1)(1 + jwR_4C_4) = jwR_2C_4R_3$$

$$R_1 + jwC_4R_4R_1 + jwL_1 + j^2w^2L_1C_4R_4 = jwC_4R_2R_3$$

$$(R_1 - w^2L_1C_4R_4) + j(wC_4R_4R_1 + wL_1) = jwC_4R_2R_3$$

Comparing the real term,

$$R_1 - w^2L_1C_4R_4 = 0$$

$$R_1 = w^2L_1C_4R_4 \tag{2.14}$$

Comparing the imaginary terms,

$$wC_4R_4R_1 + wL_1 = wC_4R_2R_3$$

$$C_4R_4R_1 + L_1 = C_4R_2R_3$$

$$L_1 = C_4R_2R_3 - C_4R_4R_1 \qquad (2.15)$$

Substituting the value of R_1 fro eqn. 2.14 into eqn. 2.15, we have,

$$L_1 = C_4R_2R_3 - C_4R_4 \times w^2L_1C_4R_4$$

$$L_1 = C_4R_2R_3 - w^2L_1C_4{}^2R_4{}^2$$

$$L_1(1 + w^2L_1C_4{}^2R_4{}^2) = C_4R_2R_3$$

$$L_1 = \frac{C_4R_2R_3}{1 + w^2L_1C_4{}^2R_4{}^2} \qquad (2.16)$$

Substituting the value of L_1 in eqn. 2.14 , we have

$$R_1 = \frac{w^2C_4{}^2R_2R_3R_4}{1 + w^2C_4{}^2R_4{}^2} \qquad (2.17)$$

$$Q = \frac{wL_1}{R_1} = \frac{w \times C_4R_2R_3}{1 + w^2C_4{}^2R_4{}^2} \times \frac{1 + w^2C_4{}^2R_4{}^2}{w^2C_4{}^2R_4R_2R_3}$$

$$Q = \frac{1}{wC_4R_4} \qquad (2.18)$$

Advantages

- ✓ Fixed capacitor is cheaper than variable capacitor.
- ✓ This bridge is best suitable for measuring high value of Q-factor.

Disadvantages

- ✓ Equations of L1and R1 are complicated.
- ✓ Measurements of R_1 and L_1 require the value of frequency.
- ✓ This bridge cannot be used for measuring low Q- factor.

2.3.4 Owen's bridge

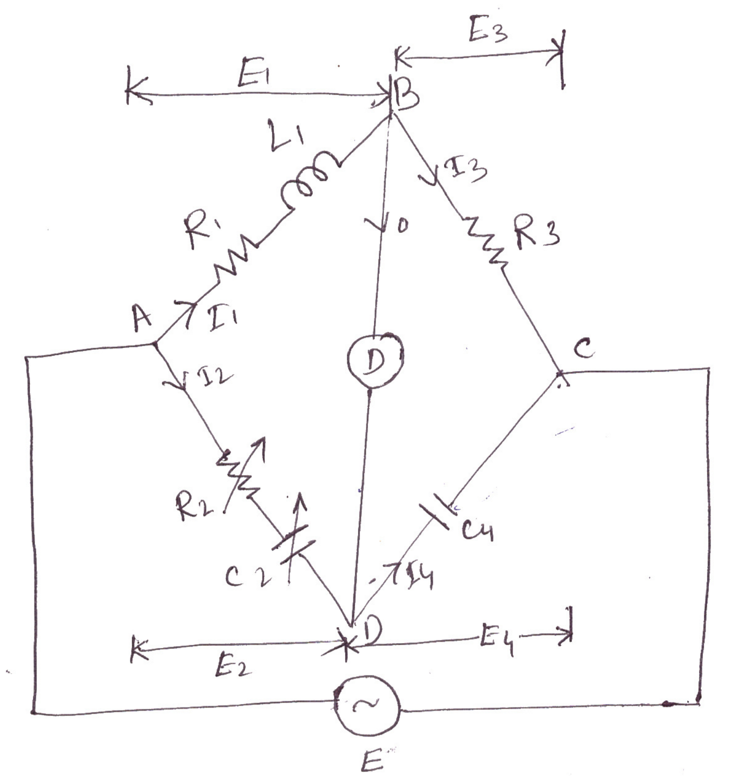


Fig 2.8 Owen's bridge

- ➢ $E_1 = I_1R_1 + jI_1X_1$
- ➢ I_4 leads E_4 by 90^0

- $\dot{E} = \dot{E}_1 + \dot{E}_3$
- $\dot{E}_2 = I_2 R_2 + \dfrac{I_2}{jwC_2}$

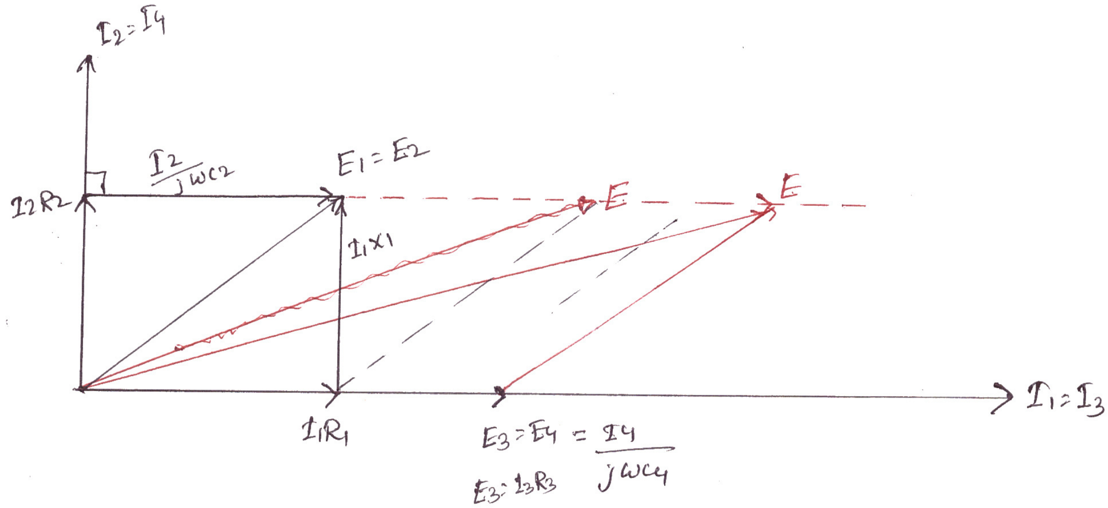


Fig 2.9 Phasor diagram of Owen's bridge

Balance condition, $\dot{Z}_1 \dot{Z}_4 = \dot{Z}_2 \dot{Z}_3$

$$Z_2 = R_2 + \frac{1}{jwC_2} = \frac{jwC_2 R_2 + 1}{jwC_2}$$

$$\therefore (R_1 + jwL_1) \times \frac{1}{jwC_4} = \frac{(1 + jwR_2 C_2) \times R_3}{jwC_2}$$

$$C_2 (R_1 + jwL_1) = R_3 C_4 (1 + jwR_2 C_2)$$

$$R_1 C_2 + jwL_1 C_2 = R_3 C_4 + jwR_2 C_2 R_3 C_4$$

Comparing real terms,

$$R_1 C_2 = R_3 C_4$$

$$R_1 = \frac{R_3 C_4}{C_2}$$

Comparing imaginary terms,

$$wL_1 C_2 = wR_2 C_2 R_3 C_4$$

$$L_1 = R_2 R_3 C_4$$

Q- factor = $\dfrac{WL_1}{R_1} = \dfrac{wR_2 R_3 C_4 C_2}{R_3 C_4}$

$$Q = wR_2 C_2$$

Advantages

- ✓ Expression for R_1 and L_1 are simple.
- ✓ R_1 and L_1 are independent of Frequency.

Disadvantages

- ✓ The Circuits used two capacitors.
- ✓ Variable capacitor is costly.
- ✓ Q-factor range is restricted.

2.3.5 Anderson's bridge

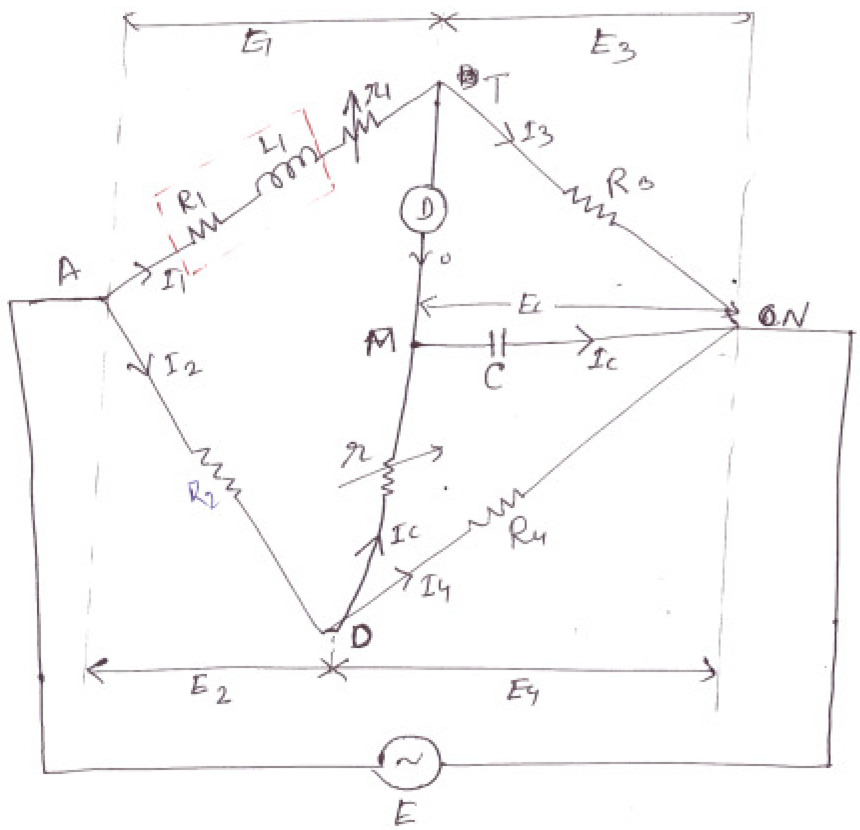


Fig 2.10 Anderson's bridge

- $\dot{E}1 = I_1(R_1 + r_1) + jI_1X_1$
- $E_3 = E_C$
- $\dot{E}4 = \dot{I}_C r + E_C$
- $I_2 = I_4 + I_C$
- $\bar{E}2 + \bar{E}4 = \bar{E}$
- $\bar{E}1 + \bar{E}3 = \bar{E}$

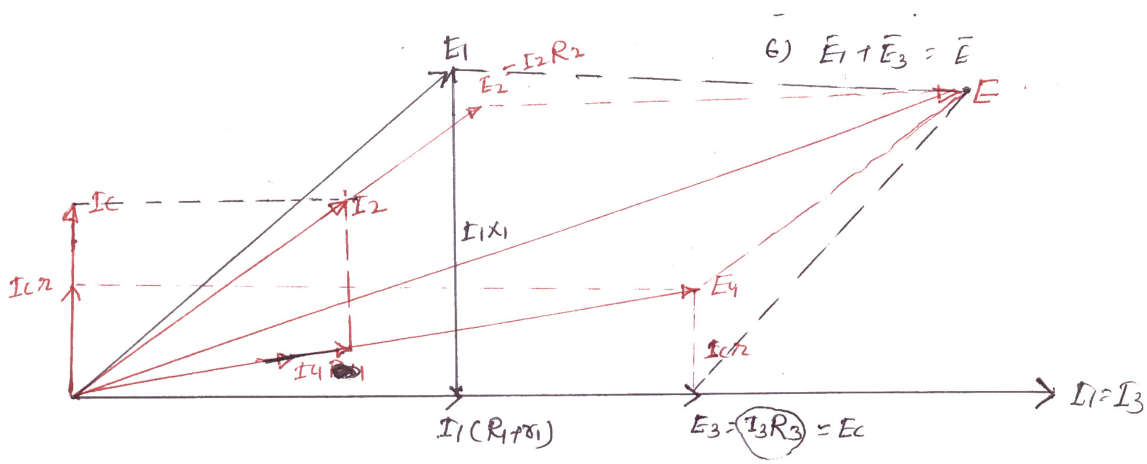


Fig 2.11 Phasor diagram of Anderson's bridge

Step-1 Take I_1 as references vector .Draw $I_1R_1^1$ in phase with I_1

$R_1^1=(R_1+r_1)$, I_1X_1 is $\perp_r$ to $I_1R_1^1$

$E_1=I_1R_1^1+jI_1X_1$

Step-2 $I_1=I_3$, E_3 is in phase with I_3, From the circuit ,

$E_3=E_C$, I_C leads E_C by 90^0

Step-3 $E_4=I_Cr+E_C$

Step-4 Draw I_4 in phase with E_4 , By KCL , $\bar{I}_2=\bar{I}_4+\bar{I}_C$

Step-5 Draw E_2 in phase with I_2

Step-6 By KVL , $\bar{E}_1+\bar{E}_3=\bar{E}$ or $\bar{E}_2+\bar{E}_4=\bar{E}$

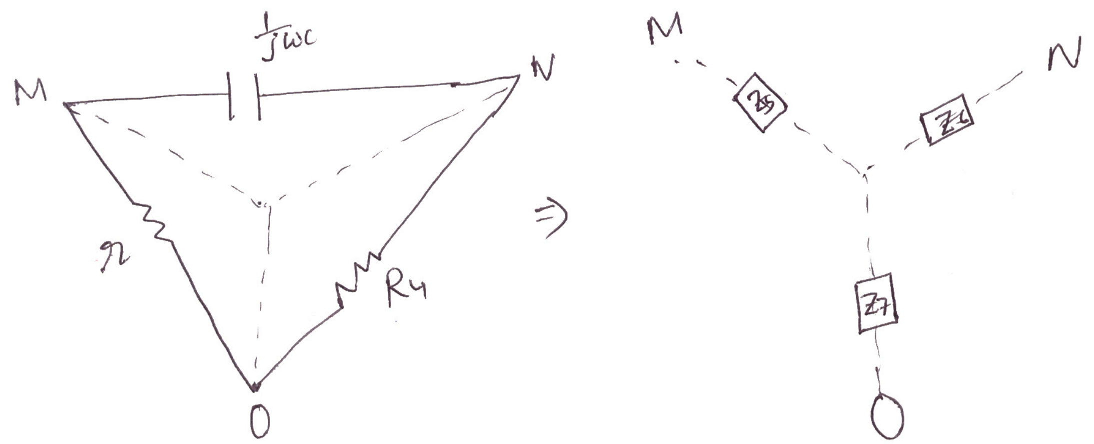


Fig 2.12 Equivalent delta to star conversion for the loop MON

$$Z_7 = \frac{R_4 \times r}{R_4 + r + \frac{1}{jwc}} = \frac{jwCR_4 r}{1 + jwC(R_4 + r)}$$

$$Z_6 = \frac{R_4 \times \frac{1}{jwC}}{R_4 + r + \frac{1}{jwc}} = \frac{R_4}{1 + jwC(R_4 + r)}$$

$$(R_1^1 + jwL_1) \times \frac{R_4}{1 + jwC(R_4 + r)} = R_3 (R_2 + \frac{jwCR_4 r}{1 + jwC(R_4 + r)})$$

$$\Rightarrow \frac{(R_1^1 + jwL_1)R_4}{1 + jwC(R_4 + r)} = R_3 \left[\frac{R_2(1 + jwC(R_4 + r)) + jwCrR_4}{1 + jwC(R_4 + r)} \right]$$

$$\Rightarrow R_1^1 R_4 + jwL_1 R_4 = R_2 R_3 + jCwR_2 R_3 (r + R_4) + jwCrR_4 R_3$$

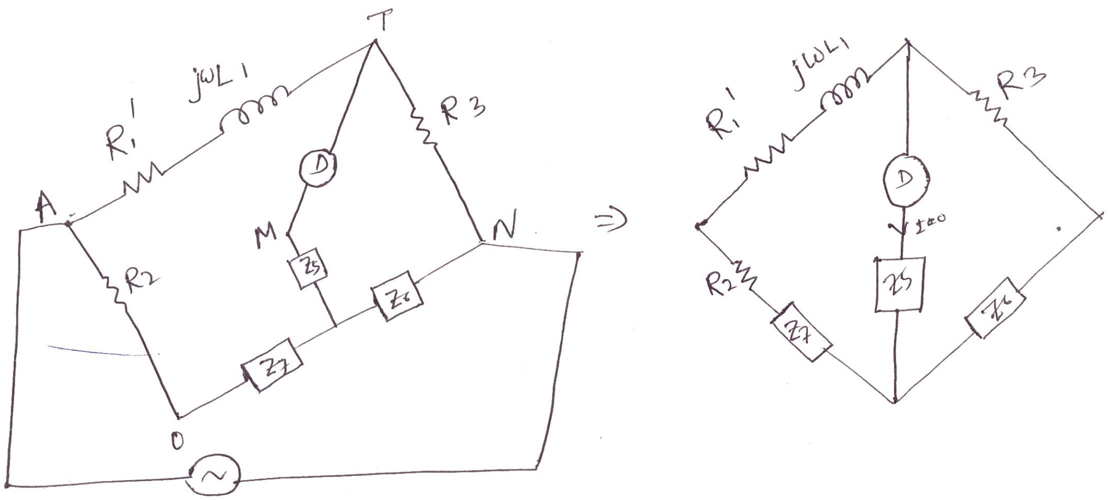


Fig 2.13 Simplified diagram of Anderson's bridge

Comparing real term,

$$R_1^1 R_4 = R_2 R_3$$

$$(R_1 + r_1)R_4 = R_2 R_3$$

$$R_1 = \frac{R_2 R_3}{R_4} - r_1$$

Comparing the imaginary term,

$$wL_1 R_4 = wCR_2 R_3 (r + R_4) + wcrR_3 R_4$$

$$L_1 = \frac{R_2 R_3 C}{R_4}(r + R_4) + R_3 rC$$

$$L_1 = R_3 C\left[\frac{R_2}{R_4}(r + R_4) + r\right]$$

Advantages

- ✓ Variable capacitor is not required.
- ✓ Inductance can be measured accurately.
- ✓ R_1 and L_1 are independent of frequency.
- ✓ Accuracy is better than other bridges.

Disadvantages

- ✓ Expression for R_1 and L_1 are complicated.
- ✓ This is not in the standard form A.C. bridge.

2.4 Measurement of capacitance and loss angle. (Dissipation factor)

2.4.1 Dissipation factors (D)

A practical capacitor is represented as the series combination of small resistance and ideal capacitance.

From the vector diagram, it can be seen that the angle between voltage and current is slightly less than 90^0. The angle 'δ' is called loss angle.

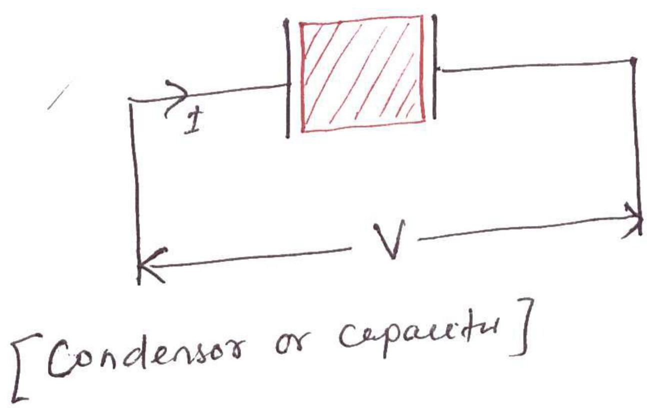


Fig 2.14 Condensor or capacitor

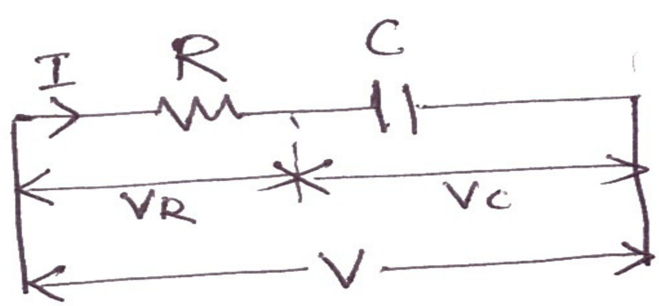


Fig 2.15 Representation of a practical capacitor

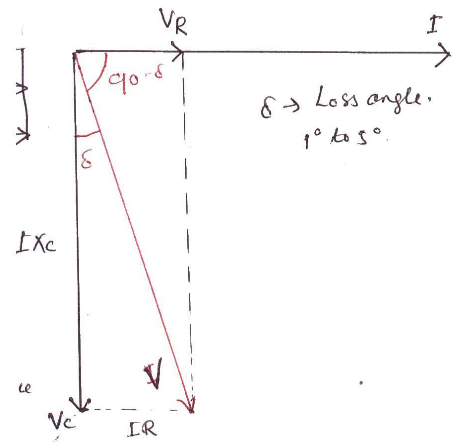


Fig 2.16 Vector diagram for a practical capacitor

A dissipation factor is defined as 'tan δ'.

$$\therefore \tan\delta = \frac{IR}{IX_C} = \frac{R}{X_C} = wCR$$

$$D = wCR$$

$$D = \frac{1}{Q}$$

$$D = \tan\delta = \frac{\sin\delta}{\cos\delta} \cong \frac{\delta}{1}$$ For small value of 'δ' in radians

$D \cong \delta \cong$ Loss Angle ('δ' must be in radian)

2.4.2 Desauty's Bridge

C_1= Unknown capacitance

At balance condition,

$$\frac{1}{jwC_1} \times R_4 = \frac{1}{jwC_2} \times R_3$$

$$\frac{R_4}{C_1} = \frac{R_3}{C_2}$$

$$\Rightarrow C_1 = \frac{R_4 C_2}{R_3}$$

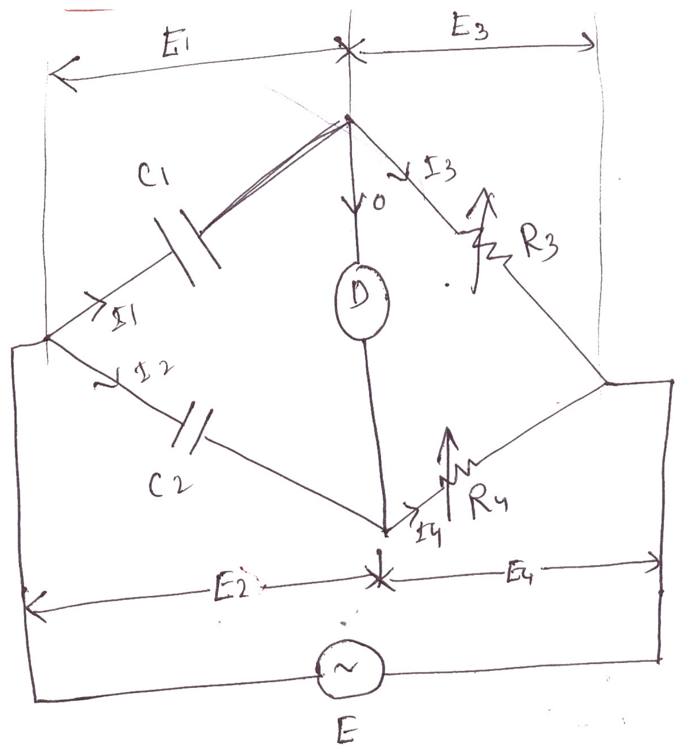


Fig 2.17 Desauty's bridge

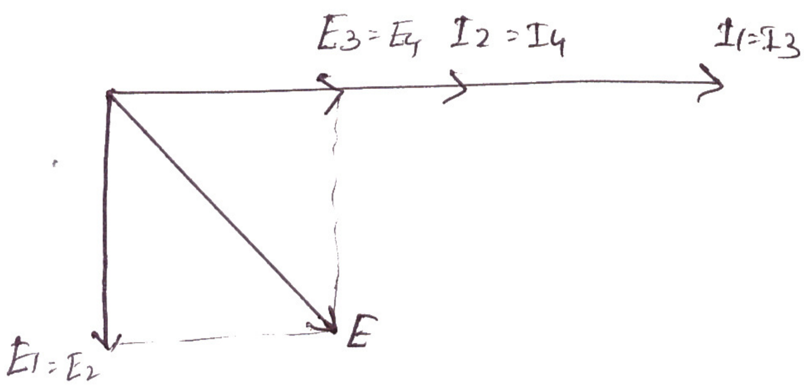


Fig 2.18 Phasor diagram of Desauty's bridge

2.4.3 Modified desauty's bridge

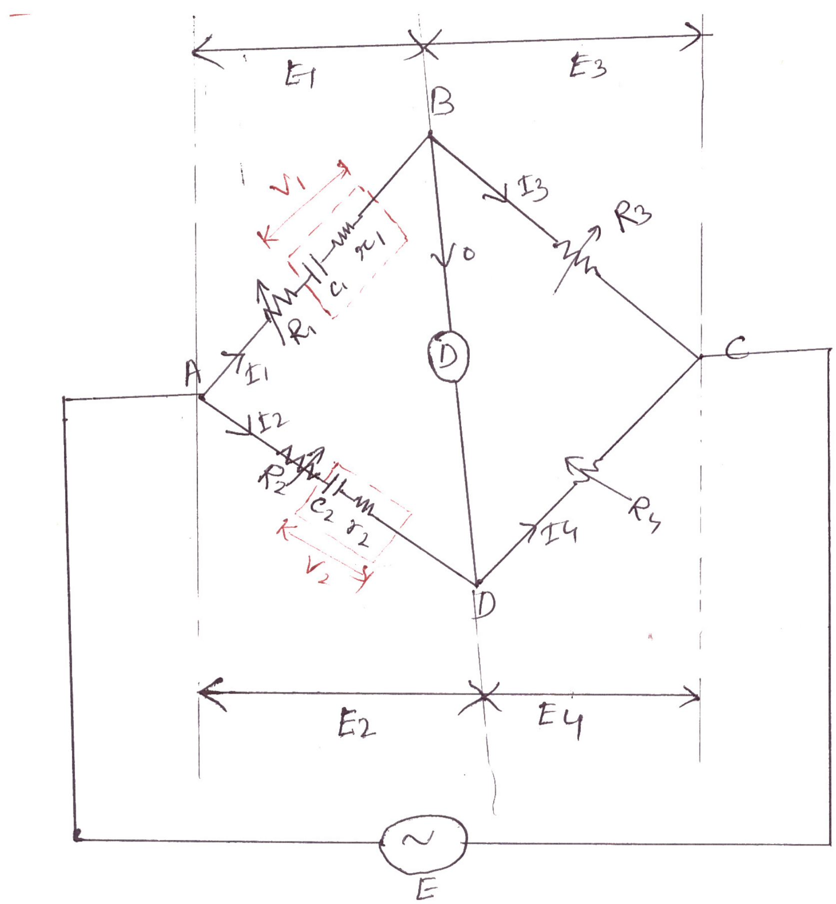


Fig 2.19 Modified Desauty's bridge

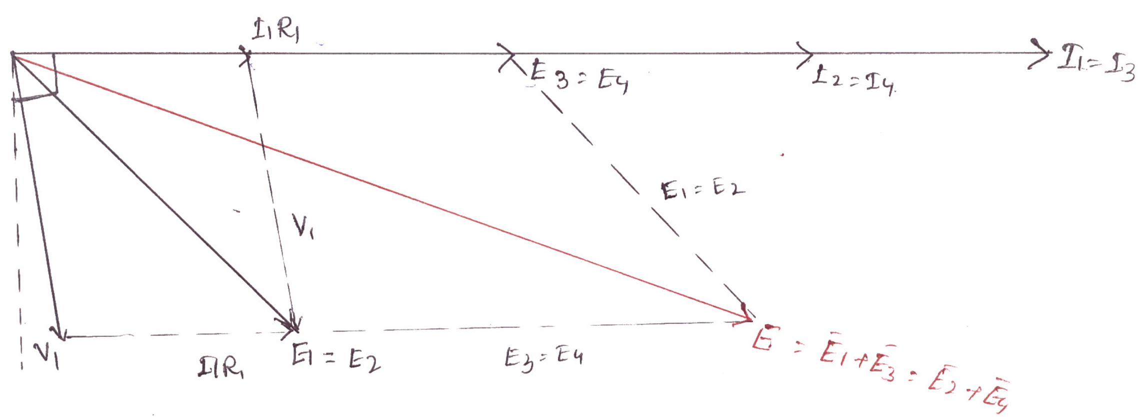


Fig 2.20 Phasor diagram of Modified Desauty's bridge

$R_1^1 = (R_1 + r_1)$

$R_2^1 = (R_2 + r_2)$

At balance condition, $(R_1^1 + \frac{1}{jwC_1})R_4 = R_3(R_2^1 + \frac{1}{jwC_2})$

$R_1^1 R_4 + \frac{R_4}{jwC_1} = R_3 R_2^1 + \frac{R_3}{jwC_2})$

Comparing the real term, $R_1^1 R_4 = R_3 R_2^1$

$R_1^1 = \frac{R_3 R_2^1}{R_4}$

$R + r_1 = \frac{(R_2 + r_2)R_3}{R_4}$

Comparing imaginary term,

$\frac{R_4}{wC_1} = \frac{R_3}{wC_2}$

$C_1 = \frac{R_4 C_2}{R_3}$

Dissipation factor $D = wC_1r_1$

Advantages

- ✓ r_1 and c_1 are independent of frequency.
- ✓ They are independent of each other.
- ✓ Source need not be pure sine wave.

2.4.4 Schering bridge

$E_1 = I_1 r_1 - jI_1 X_4$

$C_2 = C_4$= Standard capacitor (Internal resistance=0)

C_4= Variable capacitance.

C_1= Unknown capacitance.

r_1= Unknown series equivalent resistance of the capacitor.

R_3=R_4= Known resistor.

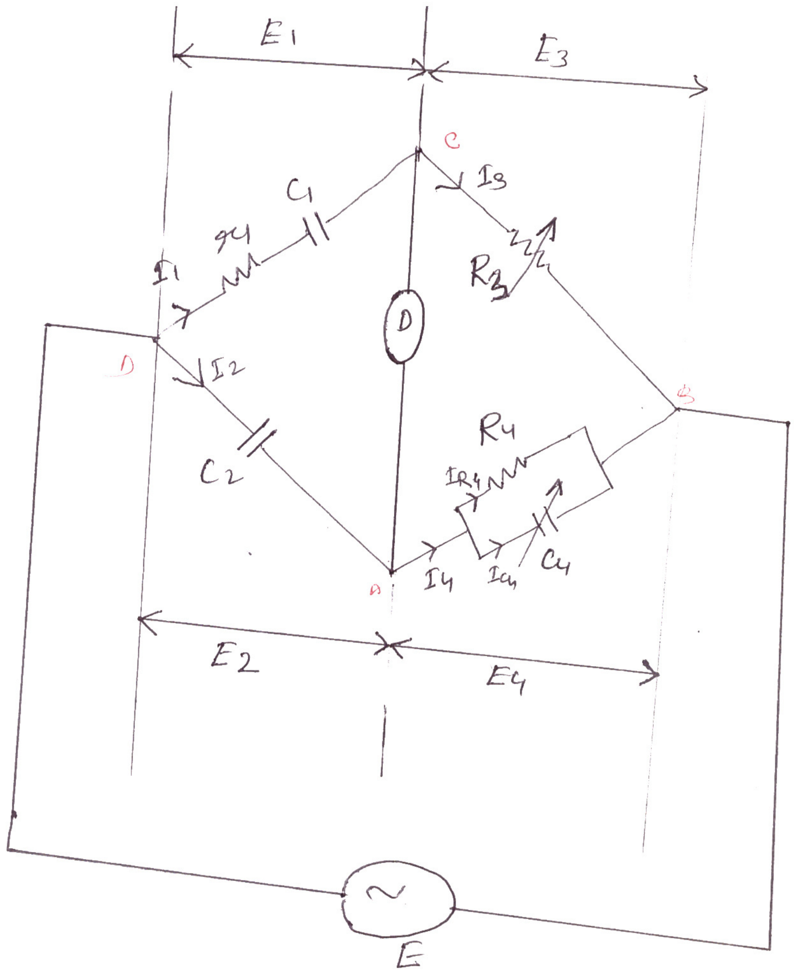


Fig 2.21 Schering bridge

$$Z_1 = r_1 + \frac{1}{jwC_1} = \frac{jwC_1 r_1 + 1}{jwC_1}$$

$$Z_4 = \frac{R_4 \times \frac{1}{jwC_4}}{R_4 + \frac{1}{jwC_4}} = \frac{R_4}{1 + jwC_4 R_4}$$

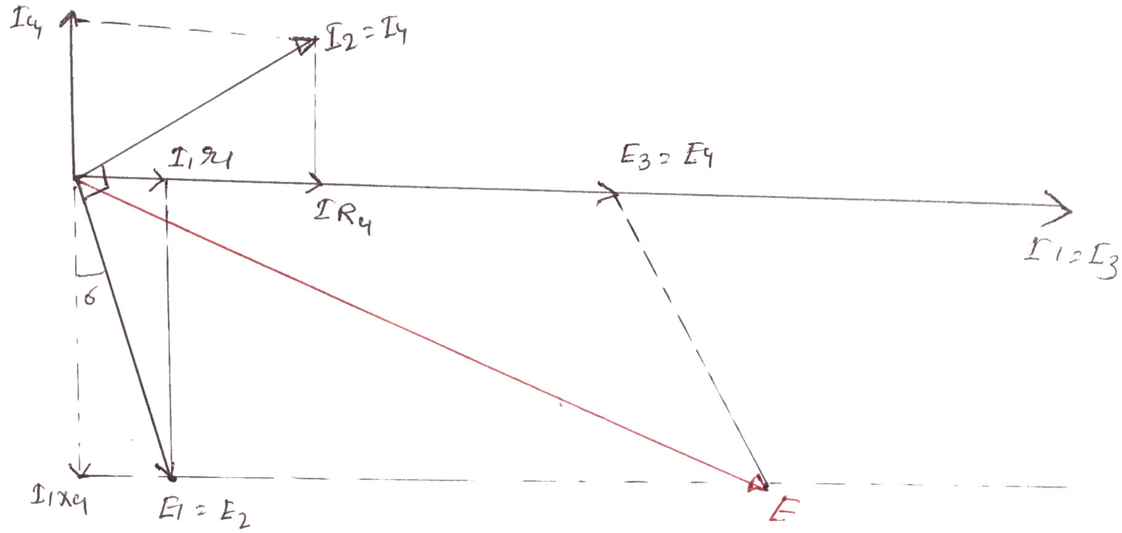


Fig 2.22 Phasor diagram of Schering bridge

At balance condition, $\dot{Z}_1 \dot{Z}_4 = \dot{Z}_2 \dot{Z}_3$

$$\frac{1 + jwC_1r_1}{jwC_1} \times \frac{R_4}{1 + jwC_4R_4} = \frac{R_3}{jwC_2}$$

$$(1 + jwC_1r_1)R_4C_2 = R_3C_1(1 + jwC_4r_4)$$

$$R_2C_2 + jwC_1r_1R_4C_2 = R_3C_1 + jwC_4R_4R_3C_1$$

Comparing the real part,

$$\therefore C_1 = \frac{R_4C_2}{R_3}$$

Comparing the imaginary part,

$$wC_1r_1R_4C_2 = wC_4R_3R_4C_1$$

$$r_1 = \frac{C_4R_3}{C_2}$$

Dissipation factor of capacitor,

$$D = wC_1 r_1 = w \times \frac{R_4 C_2}{R_3} \times \frac{C_4 R_3}{C_2}$$

$$\therefore D = wC_4 R_4$$

Advantages

- ✓ In this type of bridge, the value of capacitance can be measured accurately.
- ✓ It can measure capacitance value over a wide range.
- ✓ It can measure dissipation factor accurately.

Disadvantages

- ✓ It requires two capacitors.
- ✓ Variable standard capacitor is costly.

2.5 Measurements of frequency

2.5.1 Wein's bridge

Wein's bridge is popularly used for measurements of frequency of frequency. In this bridge, the value of all parameters are known. The source whose frequency has to measure is connected as shown in the figure.

$$Z_1 = r_1 + \frac{1}{jwC_1} = \frac{jwC_1 r_1 + 1}{jwC_1}$$

$$Z_2 = \frac{R_2}{1 + jwC_2 R_2}$$

At balance condition, $\dot{Z}_1 \dot{Z}_4 = \dot{Z}_2 \dot{Z}_3$

$$\frac{jwC_1 r_1 + 1}{jwC_1} \times R_4 = \frac{R_2}{1 + jwC_2 R_2} \times R_3$$

$$(1 + jwC_1 r_1)(1 + jwC_2 R_2) R_4 = R_2 R_3 \times jwC_1$$

$$\left[1 + jwC_2 R_2 + jwC_1 r_1 - w^2 C_1 C_2 r_1 R_2\right] = jwC_1 \frac{R_2 R_3}{R_4}$$

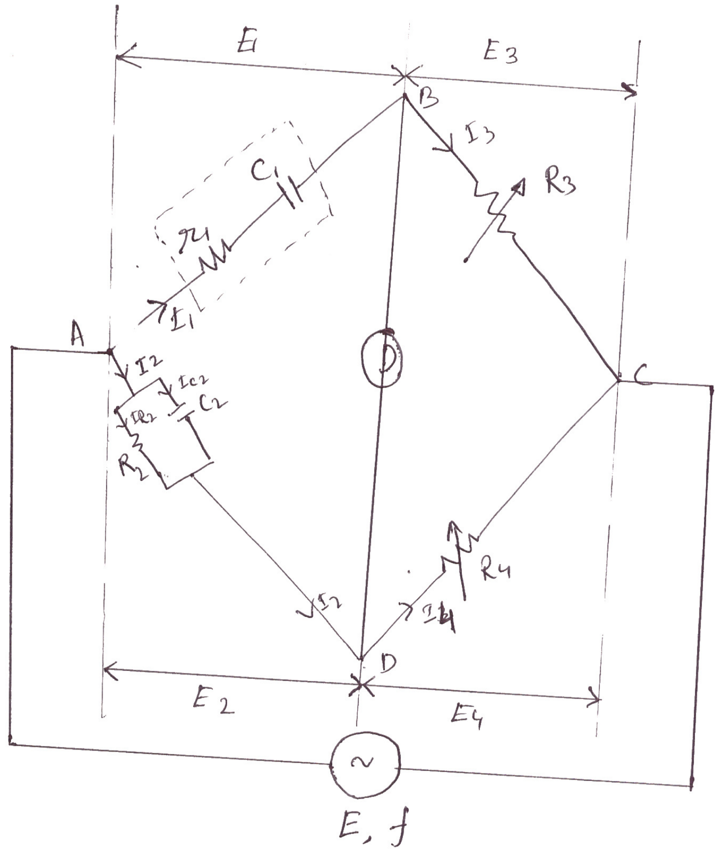


Fig 2.23 Wein's bridge

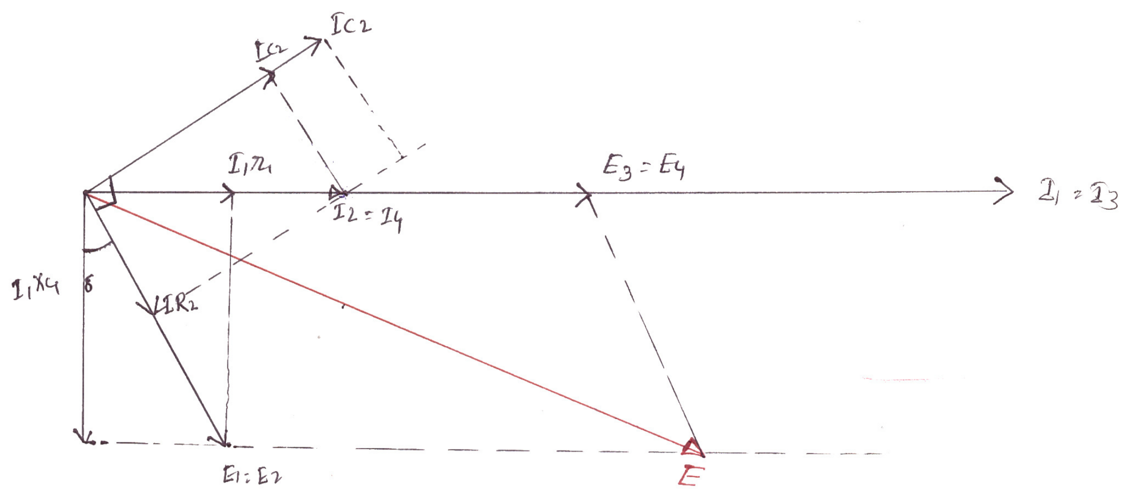


Fig 2.24 Phasor diagram of Wein's bridge

Comparing real term,

$$1 - w^2 C_1 C_2 r_1 R_2 = 0$$

$$w^2 C_1 C_2 r_1 R_2 = 1$$

$$w^2 = \frac{1}{C_1 C_2 r_1 R_2}$$

$$w = \frac{1}{\sqrt{C_1 C_2 r_1 R_2}}, \ f = \frac{1}{2\Pi\sqrt{C_1 C_2 r_1 R_2}}$$

NOTE

The above bridge can be used for measurements of capacitance. In such case, r_1 and C_1 are unknown and frequency is known. By equating real terms, we will get R_1 and C_1. Similarly by equating imaginary term, we will get another equation in terms of r_1 and C_1. It is only used for measurements of Audio frequency.

A.F=20 HZ to 20 KHZ

R.F=>> 20 KHZ

Comparing imaginary term,

$$wC_2 R_2 + wC_1 r_1 = wC_1 \frac{R_2 R_3}{R_4}$$

$$C_2 R_2 + C_1 r_1 = \frac{C_1 R_2 R_3}{R_4} \quad \ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots..(2.19)$$

$$C_1 = \frac{1}{w^2 C_2 r_1 R_2}$$

Substituting in eqn. (2.19), we have

$$C_2 R_2 + \frac{r_1}{w^2 C_2 r_1 R_2} = \frac{R_2 R_3}{R_4} C_1$$

Multiplying $\frac{R_4}{R_2 R_3}$ in both sides, we have

$$C_2 R_2 \times \frac{R_4}{R_2 R_3} + \frac{1}{w^2 C_2 R_2} \times \frac{R_4}{R_2 R_3} = C_1$$

$$C_1 = \frac{C_2 R_4}{R_3} + \frac{R_4}{w^2 C_2 R_2^2 R_3}$$

$$w^2 C_1 r_1 C_2 R_2 = 1$$

$$r_1 = \frac{1}{w^2 C_2 R_2 C_1} = \frac{1}{w^2 C_2 R_2 \left[\frac{C_2 R_4}{R_3} + \frac{R_4}{w^2 C_2 R_2^2 R_3} \right]}$$

$$= \frac{1}{\left[\frac{w^2 {C_2}^2 R_2 R_4}{R_3} + \frac{R_4}{R_2 R_3} \right]}$$

$$\therefore r_1 = \frac{1}{\frac{R_3}{R_4}\left[w^2 C_2^2 R_2 + \frac{1}{R_2} \right]}$$

$$\therefore r_1 = \frac{R_3}{R_4}\left[\frac{1}{(w^2 C_2^2 R_2 + \frac{1}{R_2})} \right]$$

2.5.2 High Voltage Schering Bridge

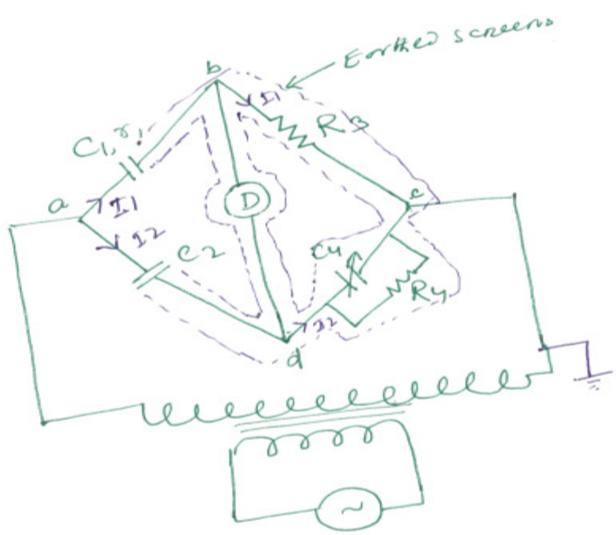


Fig 2.25 High Voltage Schering bridge

(1) The high voltage supply is obtained from a transformer usually at 50 HZ.

2.6 Wagner earthing device:

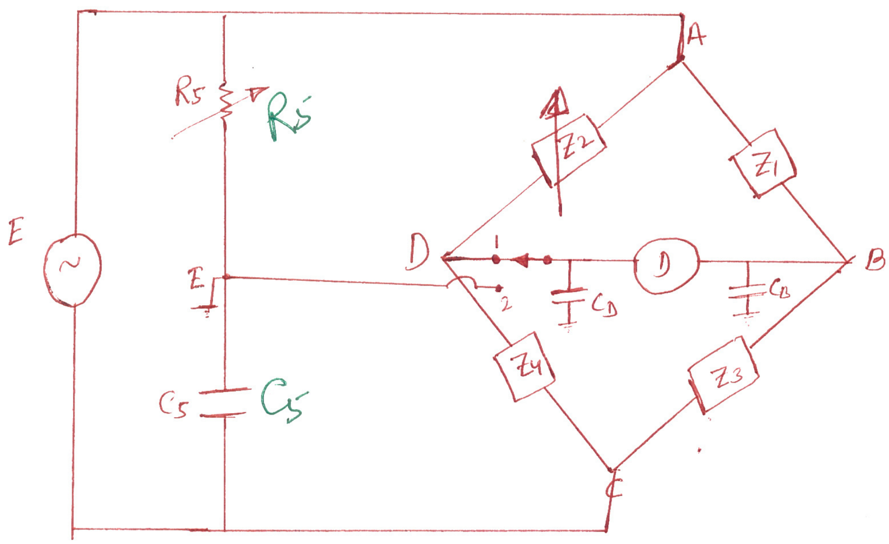


Fig 2.26 Wagner Earthing device

Wagner earthing consists of 'R' and 'C' in series. The stray capacitance at node 'B' and 'D' are C_B, C_D respectively. These Stray capacitances produced error in the measurements of 'L' and 'C'. These error will predominant at high frequency. The error due to this capacitance can be eliminated using wagner earthing arm.

Close the change over switch to the position (1) and obtained balanced. Now change the switch to position (2) and obtained balance. This process has to repeat until balance is achieved in both the position. In this condition the potential difference across each capacitor is zero. Current drawn by this is zero. Therefore they do not have any effect on the measurements.

What are the sources of error in the bridge measurements?

- ✓ Error due to stray capacitance and inductance.
- ✓ Due to external field.
- ✓ Leakage error: poor insulation between various parts of bridge can produced this error.
- ✓ Eddy current error.
- ✓ Frequency error.

- ✓ Waveform error (due to harmonics)
- ✓ Residual error: small inductance and small capacitance of the resistor produce this error.

Precaution

- ✓ The load inductance is eliminated by twisting the connecting the connecting lead.
- ✓ In the case of capacitive bridge, the connecting lead are kept apart.($\because C = \frac{A \in 0 \in r}{d}$)
- ✓ In the case of inductive bridge, the various arm are magnetically screen.
- ✓ In the case of capacitive bridge, the various arm are electro statically screen to reduced the stray capacitance between various arm.
- ✓ To avoid the problem of spike, an inter bridge transformer is used in between the source and bridge.
- ✓ The stray capacitance between the ends of detector to the ground, cause difficulty in balancing as well as error in measurements. To avoid this problem, we use wagner earthing device.

2.7 Ballastic galvanometer

This is a sophisticated instrument. This works on the principle of PMMC meter. The only difference is the type of suspension is used for this meter. Lamp and glass scale method is used to obtain the deflection. A small mirror is attached to the moving system. Phosphorous bronze wire is used for suspension.

When the D.C. voltage is applied to the terminals of moving coil, current flows through it. When a current carrying coil kept in the magnetic field, produced by permanent magnet, it experiences a force. The coil deflects and mirror deflects. The light spot on the glass scale also move. This deflection is proportional to the current through the coil.

$i = \frac{Q}{t}, Q = it = \int idt$

$\theta \propto Q$, deflection $\propto$ Charge

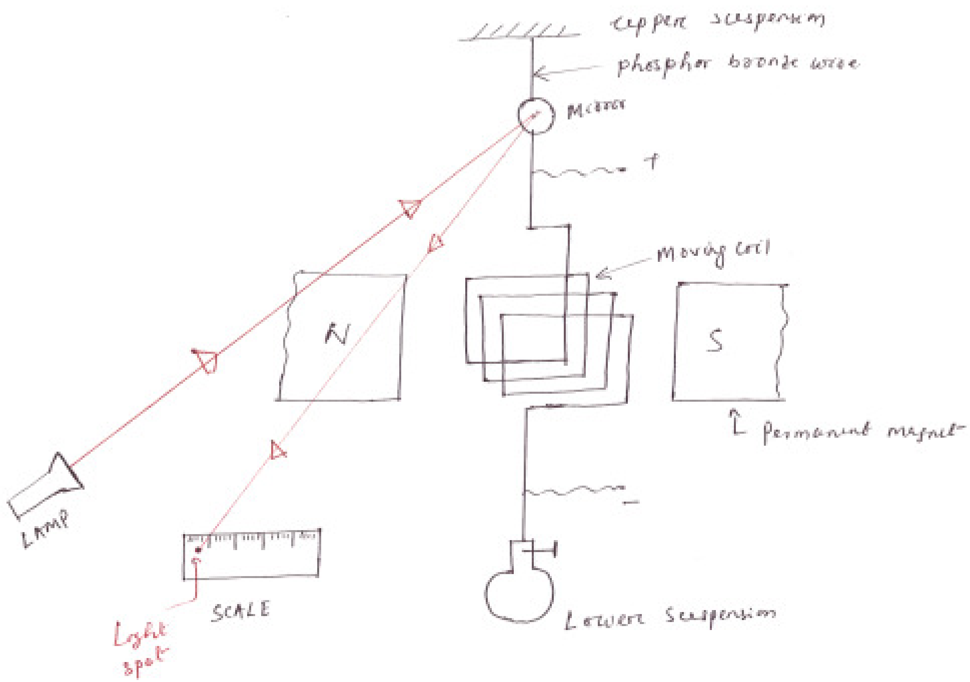


Fig 2.27 Ballastic galvanometer

2.8 Measurements of flux and flux density (Method of reversal)

D.C. voltage is applied to the electromagnet through a variable resistance R_1 and a reversing switch. The voltage applied to the toroid can be reversed by changing the switch from position 2 to position '1'. Let the switch be in position '2' initially. A constant current flows through the toroid and a constant flux is established in the core of the magnet.

A search coil of few turns is provided on the toroid. The B.G. is connected to the search coil through a current limiting resistance. When it is required to measure the flux, the switch is changed from position '2' to position '1'. Hence the flux reduced to zero and it starts increasing in the reverse direction. The flux goes from + ϕ to - ϕ, in time 't' second. An emf is induced in the search coil, science the flux changes with time. This emf circulates a current through R_2 and B.G. The meter deflects. The switch is normally closed. It is opened when it is required to take the reading.

2.8.1 Plotting the BH curve

The curve drawn with the current on the X-axis and the flux on the Y-axis, is called magnetization characteristics. The shape of B-H curve is similar to shape of magnetization characteristics. The residual magnetism present in the specimen can be removed as follows.

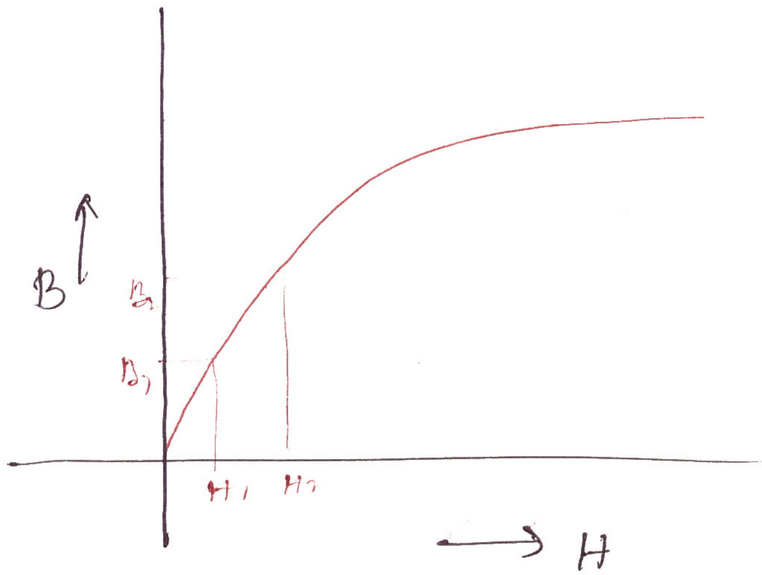


Fig 2.28 BH curve

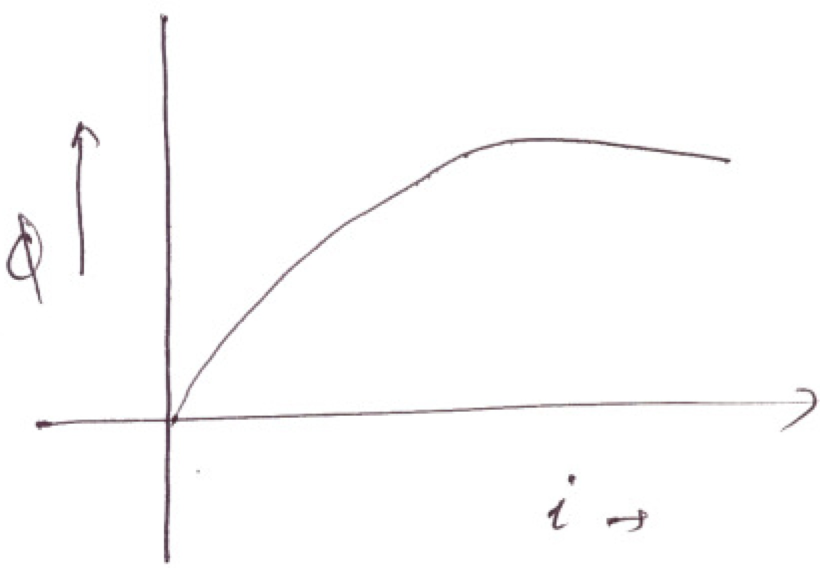


Fig 2.29 Magnetization characteristics

Close the switch ‘S_2’ to protect the galvanometer, from high current. Change the switch S1 from position ‘1’ to ‘2’ and vice versa for several times.

To start with the resistance ‘R_1’ is kept at maximum resistance position. For a particular value of current, the deflection of B.G. is noted. This process is repeated for various value of current. For each deflection flux can be calculated.($B = \frac{\phi}{A}$)

Magnetic field intensity value for various current can be calculated.().The B-H curve can be plotted by using the value of ‘B’ and ‘H’.

2.8.2 Measurements of iron loss:

Let R_P= pressure coil resistance

R_S = resistance of coil S1

E= voltage reading= Voltage induced in S_2

I= current in the pressure coil

V_P= Voltage applied to wattmeter pressure coil.

W= reading of wattmeter corresponding voltage V

W_1= reading of wattmeter corresponding voltage E

$$W \rightarrow V \qquad \frac{W_1}{W} = \frac{E}{V} \Rightarrow W_1 = \frac{E \times W}{V}$$
$$W_1 \rightarrow E_P$$

W1=Total loss=Iron loss+ Cupper loss.

The above circuit is similar to no load test of transformer.

In the case of no load test the reading of wattmeter is approximately equal to iron loss. Iron loss depends on the emf induced in the winding. Science emf is directly proportional to flux. The voltage applied to the pressure coil is V. The corresponding of wattmeter is ‘W’. The iron loss corresponding E is $E = \frac{WE}{V}$. The reading of the wattmeter includes the losses in the pressure coil and copper loss of the winding S1. These loses have to be subtracted to get the actual iron loss.

2.9 Galvanometers

D-Arsonval Galvanometer

Vibration Galvanometer

Ballistic C

2.9.1 D-arsonval galvanometer (d.c. galvanometer)

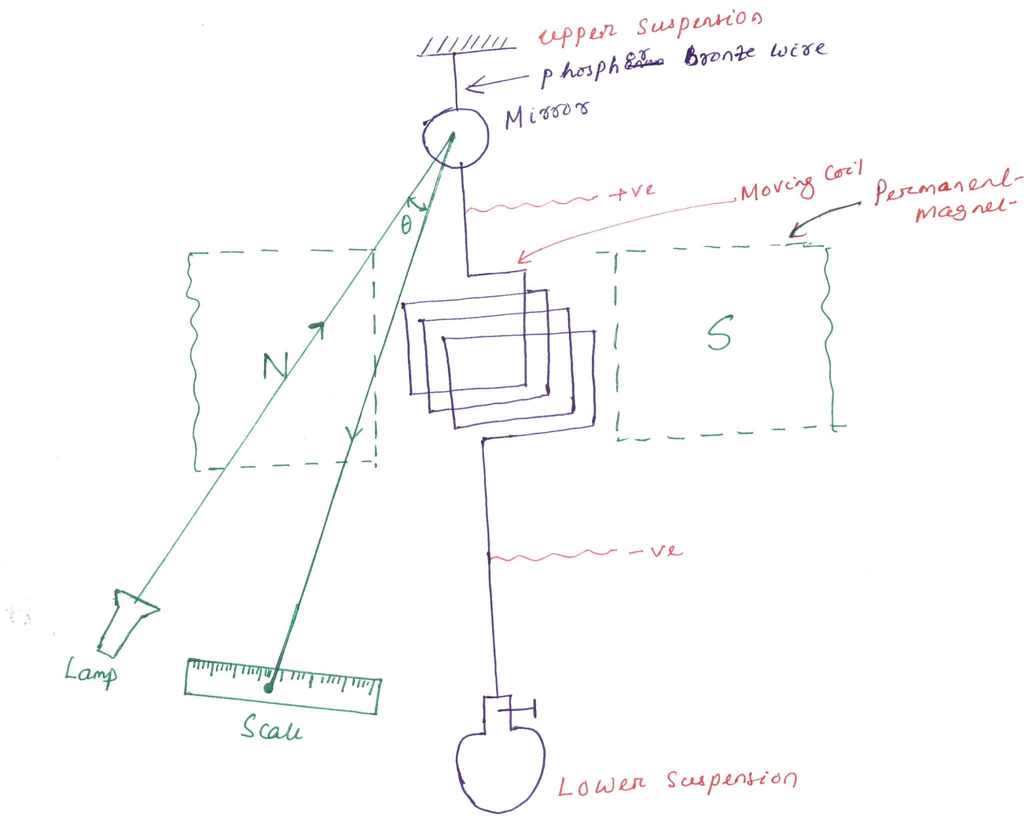


Fig 2.30 D-Arsonval Galvanometer

Galvanometer is a special type of ammeter used for measuring μA or mA. This is a sophisticated instruments. This works on the principle of PMMC meter. The only difference is the type of suspension used for this meter. It uses a sophisticated suspension called taut suspension, so that moving system has negligible weight.

Lamp and glass scale method is used to obtain the deflection. A small mirror is attached to the moving system. Phosphors bronze is used for suspension.

When D.C. voltage is applied to the terminal of moving coil, current flows through it. When current carrying coil is kept in the magnetic field produced by P.M. , it experiences a force. The light spot on the glass scale also move. This deflection is proportional to the current through the coil. This instrument can be used only with D.C. like PMMC meter.

The deflecting Torque,

T_D=BINA

T_D=GI, Where G=BAN

$T_C=K_S\theta=S\theta$

At balance, $T_C=T_D \Rightarrow S\theta=GI$

$$\therefore \theta = \frac{GI}{S}$$

Where G= Displacements constant of Galvanometer

S=Spring constant

2.9.2 Vibration Galvanometer (A.C. Galvanometer)

The construction of this galvanometer is similar to the PMMC instrument except for the moving system. The moving coil is suspended using two ivory bridge pieces. The tension of the system can be varied by rotating the screw provided at the top suspension. The natural frequency can be varied by varying the tension wire of the screw or varying the distance between ivory bridge piece.

When A.C. current is passed through coil an alternating torque or vibration is produced. This vibration is maximum if the natural frequency of moving system coincide with supply frequency. Vibration is maximum, science resonance takes place. When the coil is vibrating , the mirror oscillates and the dot moves back and front. This appears as a line on the glass scale. Vibration galvanometer is used for null deflection of a dot appears on the scale. If the bridge is unbalanced, a line appears on the scale

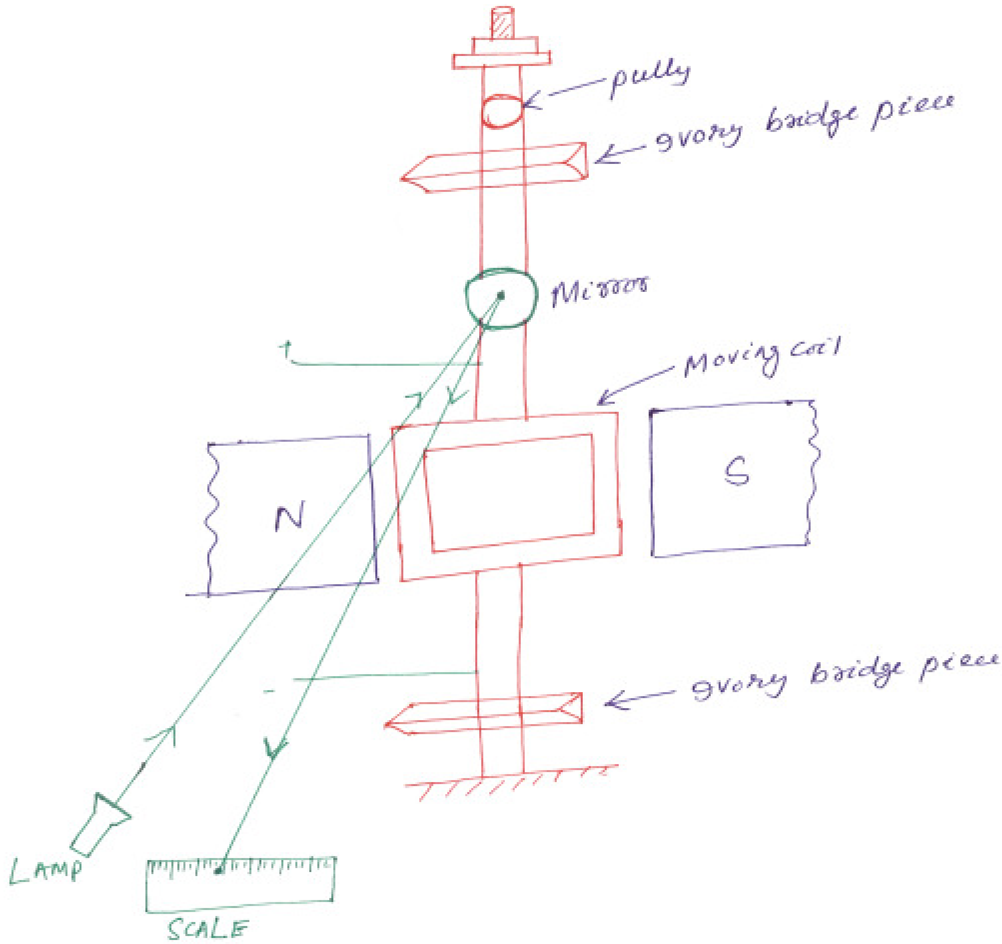


Fig 2.31 Vibration Galvanometer

Example 2.2-In a low- Voltage Schering bridge designed for the measurement of permittivity, the branch 'ab' consists of two electrodes between which the specimen under test may be inserted, arm 'bc' is a non-reactive resistor R_3 in parallel with a standard capacitor C_3, arm CD is a non-reactive resistor R_4 in parallel with a standard capacitor C_4, arm 'da' is a standard air capacitor of capacitance C_2. Without the specimen between the electrode, balance is obtained with following values , C_3=C_4=120 pF, C_2=150 pF, R_3=R_4=5000Ω.With the specimen inserted, these values become C_3=200 pF,C_4=1000 pF,C_2=900 pF and R_3=R_4=5000Ω. In such test w=5000 rad/sec. Find the relative permittivity of the specimen?

Sol: Relative permittivity$(\varepsilon_r) = \dfrac{\text{capacitance measured with given medium}}{\text{capacitance measured with air medium}}$

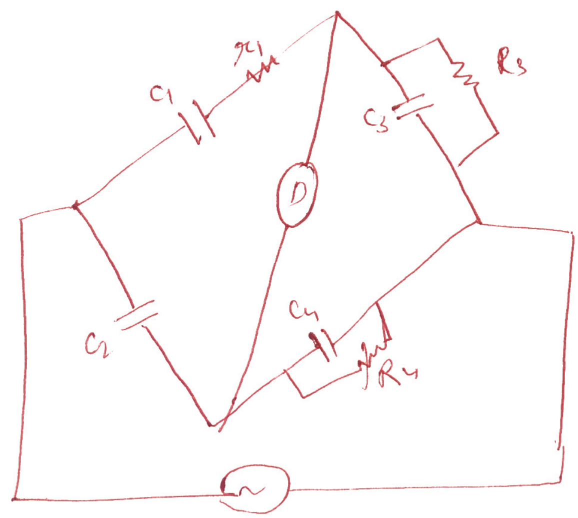


Fig 2.32 Schering bridge

$$C_1 = C_2(\frac{R_4}{R_3})$$

Let capacitance value C_0, when without specimen dielectric.

Let the capacitance value C_S when with the specimen dielectric.

$$C_0 = C_2(\frac{R_4}{R_3}) = 150 \times \frac{5000}{5000} = 150 pF$$

$$C_S = C_2(\frac{R_4}{R_3}) = 900 \times \frac{5000}{5000} = 900 pF$$

$$\varepsilon_r = \frac{C_S}{C_0} = \frac{900}{150} = 6$$

Example 2.3- A specimen of iron stamping weighting 10 kg and having a area of 16.8 cm^2 is tested by an episten square. Each of the two winding S_1 and S_2 have 515 turns. A.C. voltage of 50 HZ frequency is given to the primary. The current in the primary is 0.35 A. A voltmeter connected to S_2 indicates 250 V. Resistance of S_1 and S_2 each equal to 40 Ω. Resistance of pressure coil is 80 kΩ. Calculate maximum flux density in the specimen and iron loss/kg if the wattmeter indicates 80 watt?

Sol^n- $E = 4.44 f\phi_m N$

$$B_m = \frac{E}{4.44 fAN} = 1.3 wb/m^2$$

Iron loss= $W(1+\frac{R_S}{R_P}) - \frac{E^2}{(R_S + R_P)}$

$$= 80(1+\frac{40}{80\times10^3}) - \frac{250^2}{(40+80\times10^3)} = 79.26 watt$$

Iron loss/ kg=79.26/10=7.926 w/kg.

8. Resolution

Resolution is the smallest detectable incremental change of input parameter that can be detected in the output signal. Resolution can be expressed either as a proportion of the full-scale reading or in absolute terms. For example, if a LVDT sensor measures a displacement up to 20 mm and it provides an output as a number between 1 and 100 then the resolution of the sensor device is 0.2 mm.

9. Stability

Stability is the ability of a sensor device to give same output when used to measure a constant input over a period of time. The term 'drift' is used to indicate the change in output that occurs over a period of time. It is expressed as the percentage of full range output.

10.Dead band/time

The dead band or dead space of a transducer is the range of input values for which there is no output. The dead time of a sensor device is the time duration from the application of an input until the output begins to respond or change.

11.Repeatability

It specifies the ability of a sensor to give same output for repeated applications of same input value. It is usually expressed as a percentage of the full range output:

Repeatability = (maximum – minimum values given) X 100 / full range (2.1.2)

12.Response time

Response time describes the speed of change in the output on a step-wise change of the measurand. It is always specified with an indication of input step and the output range for which the response time is defined.

Classification of sensors

Sensors can be classified into various groups according to the factors such as measurand, application fields, conversion principle, energy domain of the measurand and thermodynamic considerations. These general classifications of sensors are well described in the references [2, 3].

Detail classification of sensors in view of their applications in manufacturing is as follows.

A. Displacement, position and proximity sensors
    - Potentiometer
    - Strain-gauged element
    - Capacitive element
    - Differential transformers
    - Eddy current proximity sensors
    - Inductive proximity switch
    - Optical encoders
    - Pneumatic sensors
    - Proximity switches (magnetic)
    - Hall effect sensors

B. Velocity and motion
    - Incremental encoder
    - Tachogenerator
    - Pyroelectric sensors

C. Force
    - Strain gauge load cell

D. Fluid pressure
    - Diaphragm pressure gauge
    - Capsules, bellows, pressure tubes
    - Piezoelectric sensors
    - Tactile sensor
    -

E. Liquid flow
    - Orifice plate
    - Turbine meter

F. Liquid level
    - Floats
    - Differential pressure

G. Temperature
    - Bimetallic strips
    - Resistance temperature detectors
    - Thermistors
    - Thermo-diodes and transistors
    - Thermocouples
    - Light sensors
    - Photo diodes
    - Photo resistors

Displacement and position sensors

Displacement sensors are basically used for the measurement of movement of an object. Position sensors are employed to determine the position of an object in relation to some reference point.

Proximity sensors are a type of position sensor and are used to trace when an object has moved with in particular critical distance of a transducer.

Displacement sensors

1. Potentiometer Sensors

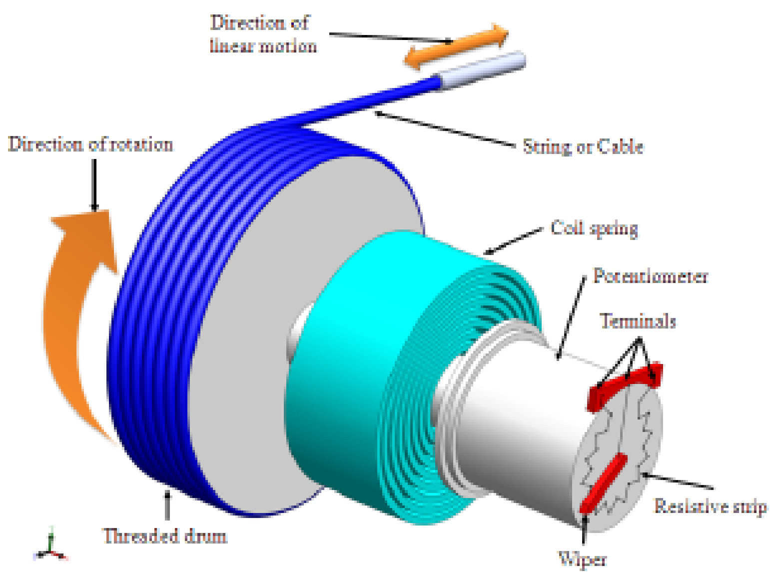


Figure 2.2.1 Schematic of a potentiometer sensor for measurement of linear displacement

Figure 2.2.1 shows the construction of a rotary type potentiometer sensor employed to measure the linear displacement. The potentiometer can be of linear or angular type. It works on the principle of conversion of mechanical displacement into an electrical signal. The sensor has a resistive element and a sliding contact (wiper). The slider moves along this conductive body, acting as a movable electric contact.

The object of whose displacement is to be measured is connected to the slider by using

- a rotating shaft (for angular displacement)
- a moving rod (for linear displacement)
- a cable that is kept stretched during operation

The resistive element is a wire wound track or conductive plastic. The track comprises of large number of closely packed turns of a resistive wire. Conductive plastic is made up of plastic resin embedded with the carbon powder. Wire wound track has a resolution of the order of ± 0.01 % while the conductive plastic may have the resolution of about 0.1 μm.

During the sensing operation, a voltage V_s is applied across the resistive element. A voltage divider circuit is formed when slider comes into contact with the wire. The output voltage (V_A) is measured as shown in the figure 2.2.2. The output voltage is proportional to the displacement of the slider over the wire. Then the output parameter displacement is calibrated against the output voltage V_A.

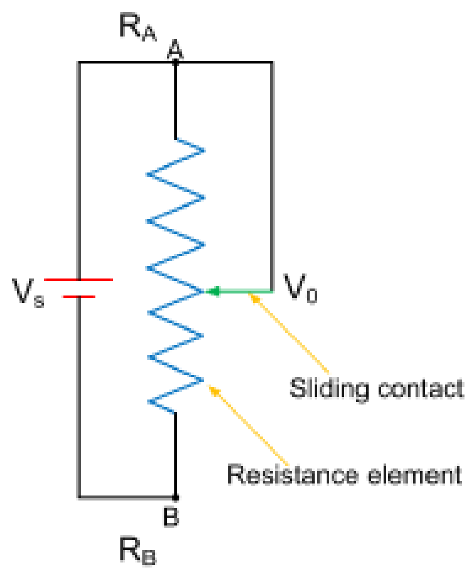


Figure 2.2.2 Potentiometer: electric circuit

$V_A = I\,R_A$ (2.2.1)

But $I = V_S / (R_A + R_B)$ (2.2.2)

Therefore $V_A = V_S\,R_A / (R_A + R_B)$ (2.2.3)

As we know that $R = \rho\,L / A$, where ρ is electrical resistivity, L is length of resistor and A is area of cross section

$V_A = V_S\,L_A / (L_A + L_B)$ (2.2.4)

Applications of potentiometer

These sensors are primarily used in the control systems with a feedback loop to ensure that the moving member or component reaches its commanded position.

These are typically used on machine-tool controls, elevators, liquid-level assemblies, forklift trucks, automobile throttle controls. In manufacturing, these are used in control of injection molding machines, woodworking machinery, printing, spraying, robotics, etc. These are also used in computer-controlled monitoring of sports equipment.

2. *Strain Gauges*

The strain in an element is a ratio of change in length in the direction of applied load to the original length of an element. The strain changes the resistance *R* of the element. Therefore, we can say,

$\Delta R/R \propto \varepsilon$;

$$\Delta R/R = G\,\varepsilon \qquad (2.2.5)$$

where G is the constant of proportionality and is called as gauge factor. In general, the value of G is considered in between 2 to 4 and the resistances are taken of the order of 100 Ω.

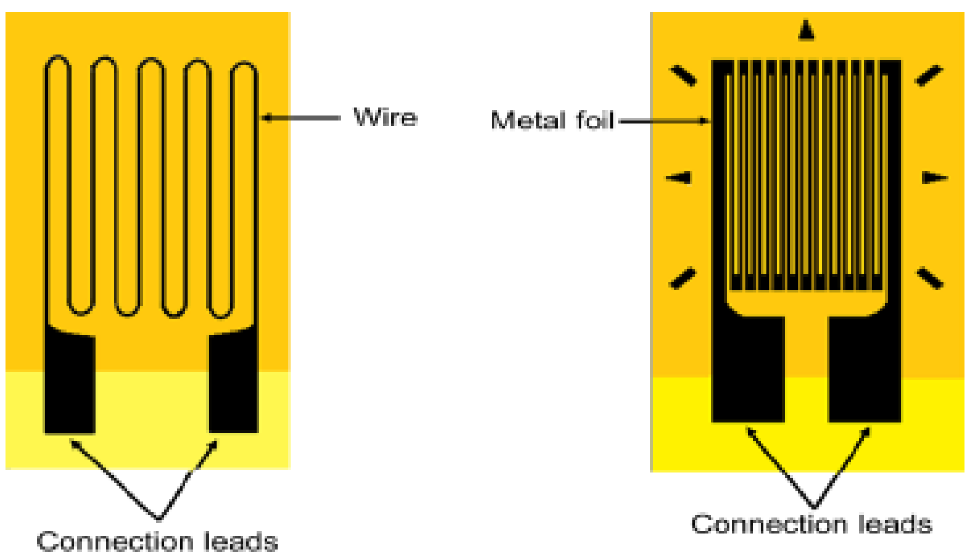


Figure 2.2.3 A pattern of resistive foils

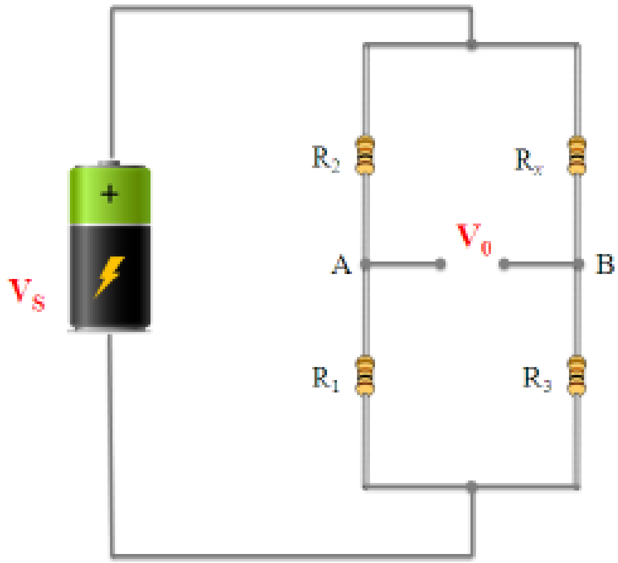


Figure 2.2.4 Wheatstone's bridge

Resistance strain gauge follows the principle of change in resistance as per the equation 2.2.5. It comprises of a pattern of resistive foil arranged as shown in Figure 2.2.3. These foils are made of Constantan alloy (copper-nickel 55-45% alloy) and are bonded to a backing material plastic (ployimide), epoxy or glass fiber reinforced epoxy. The strain gauges are secured to the workpiece by using epoxy or Cyanoacrylate cement Eastman 910 SL. As the workpiece undergoes change in its shape due to external loading, the resistance of strain gauge element changes. This change in resistance can be detected by a using a Wheatstone's resistance bridge as shown in Figure 2.2.4. In the balanced bridge we can have a relation,

$$R_2/ R_1 = R_x / R_3 \qquad (2.2.6)$$

where R_x is resistance of strain gauge element, R_2 is balancing/adjustable resistor, R_1 and R_3 are known constant value resistors. The measured deformation or displacement by the stain gauge is calibrated against change in resistance of adjustable resistor R_2 which makes the voltage across nodes A and B equal to zero.

Applications of strain gauges

Strain gauges are widely used in experimental stress analysis and diagnosis on machines and failure analysis. They are basically used for multi-axial stress fatigue testing, proof testing, residual stress and vibration measurement, torque measurement, bending and deflection measurement, compression and tension measurement and strain measurement.

Strain gauges are primarily used as sensors for machine tools and safety in automotives. In particular, they are employed for force measurement in machine tools, hydraulic or pneumatic press and as impact sensors in aerospace vehicles.

3. *Capacitive element based sensor*

Capacitive sensor is of non-contact type sensor and is primarily used to measure the linear displacements from few millimeters to hundreds of millimeters. It comprises of three plates, with the upper pair forming one capacitor and the lower pair another. The linear displacement might take in two forms:

a. one of the plates is moved by the displacement so that the plate separation changes
b. area of overlap changes due to the displacement.

Figure 2.2.5 shows the schematic of three-plate capacitive element sensor and displacement measurement of a mechanical element connected to the plate 2.

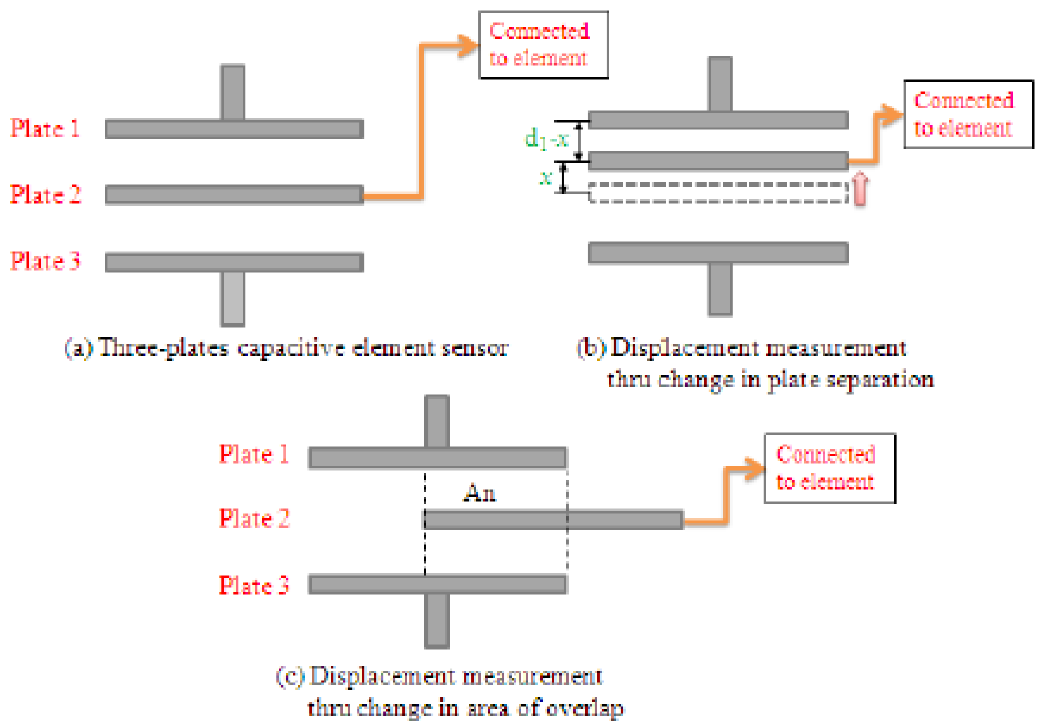


Figure 2.2.5 Displacement measurement using capacitive element sensor

The capacitance C of a parallel plate capacitor is given by,

$$C = \varepsilon_r \varepsilon_o A / d \qquad (2.2.7)$$

where ε_r is the relative permittivity of the dielectric between the plates, ε_o permittivity of free space, A area of overlap between two plates and d the plate separation.

As the central plate moves near to top plate or bottom one due to the movement of the element/workpiece of which displacement is to be measured, separation in between the plate changes. This can be given as,

$$C_1 = (\varepsilon_r \, \varepsilon_o \, A) / (d + x) \quad (2.2.8)$$

$$C_2 = (\varepsilon_r \, \varepsilon_o \, A) / (d - x) \quad (2.2.9)$$

When C1 and C2 are connected to a Wheatsone's bridge, then the resulting out-of-balance voltage would be in proportional to displacement x.

Capacitive elements can also be used as proximity sensor. The approach of the object towards the sensor plate is used for induction of change in plate separation. This changes the capacitance which is used to detect the object.

Applications of capacitive element sensors

- Feed hopper level monitoring
- Small vessel pump control
- Grease level monitoring
- Level control of liquids
- Metrology applications
    - to measure shape errors in the part being produced
    - to analyze and optimize the rotation of spindles in various machine tools such as surface grinders, lathes, milling machines, and air bearing spindles by measuring errors in the machine tools themselves
- Assembly line testing
    - to test assembled parts for uniformity, thickness or other design features
    - to detect the presence or absence of a certain component, such as glue etc.

4. Linear variable differential transformer (LVDT)

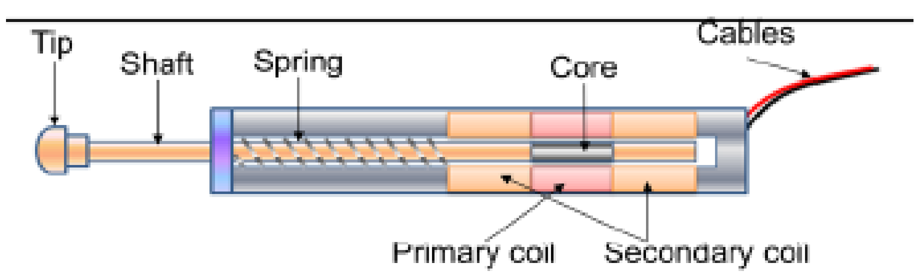


Figure 2.2.6 Construction of a LVDT sensor

Linear variable differential transformer (LVDT) is a primary transducer used for measurement of linear displacement with an input range of about ± 2 to ± 400 mm in general. It has non-linearity error ± 0.25% of full range. Figure 2.2.6 shows the construction of a LVDT sensor. It has three coils symmetrically spaced along an insulated tube. The central coil is primary coil and the other two are secondary coils. Secondary coils are connected in series in such a way that their outputs oppose each other. A magnetic core attached to the element of which displacement is to be monitored is placed inside the insulated tube.

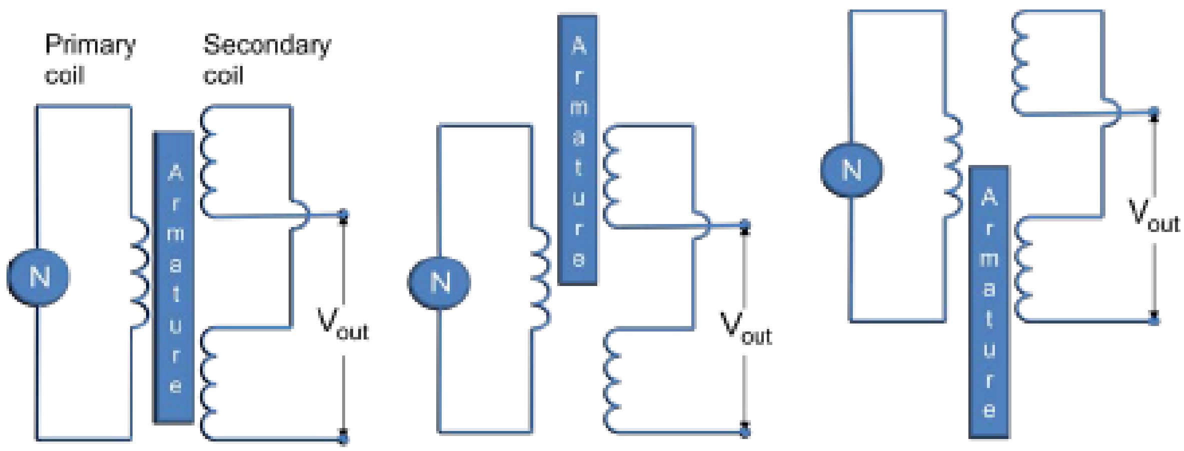


Figure 2.2.7 Working of LVDT sensor

Due to an alternating voltage input to the primary coil, alternating electro-magnetic forces (emfs) are generated in secondary coils. When the magnetic core is centrally placed with its half portion in each of the secondary coil regions then the resultant voltage is zero. If the core is displaced from the central position as shown in Figure 2.2.7, say, more in secondary coil 1 than in coil 2, then more emf is generated in one coil i.e. coil 1 than the other, and there is a resultant voltage from the coils. If the magnetic core is further displaced, then the value of resultant voltage increases in proportion with the displacement. With the help of signal processing devices such as low pass filters and demodulators, precise displacement can be measured by using LVDT sensors.

LVDT exhibits good repeatability and reproducibility. It is generally used as an absolute position sensor. Since there is no contact or sliding between the constituent elements of the sensor, it is highly reliable. These sensors are completely sealed and are widely used in Servomechanisms, automated measurement in machine tools.

A rotary variable differential transformer (RVDT) can be used for the measurement of rotation. Readers are suggested to prepare a report on principle of working and construction of RVDT sensor.

Applications of LVDT sensors

- Measurement of spool position in a wide range of servo valve applications
- To provide displacement feedback for hydraulic cylinders
- To control weight and thickness of medicinal products viz. tablets or pills
- For automatic inspection of final dimensions of products being packed for dispatch
- To measure distance between the approaching metals during Friction welding process
- To continuously monitor fluid level as part of leak detection system
- To detect the number of currency bills dispensed by an ATM

Quiz:

1. Explain the principle of working of LVDT.
2. Describe the working of RVDT with a neat sketch.
3. List the applications of potentiometer sensor in/around your home and office/university.

References

1. Boltan, W., Mechatronics: electronic control systems in mechanical and electrical engineering, Longman, Singapore, 1999.

6. Hall effect sensor

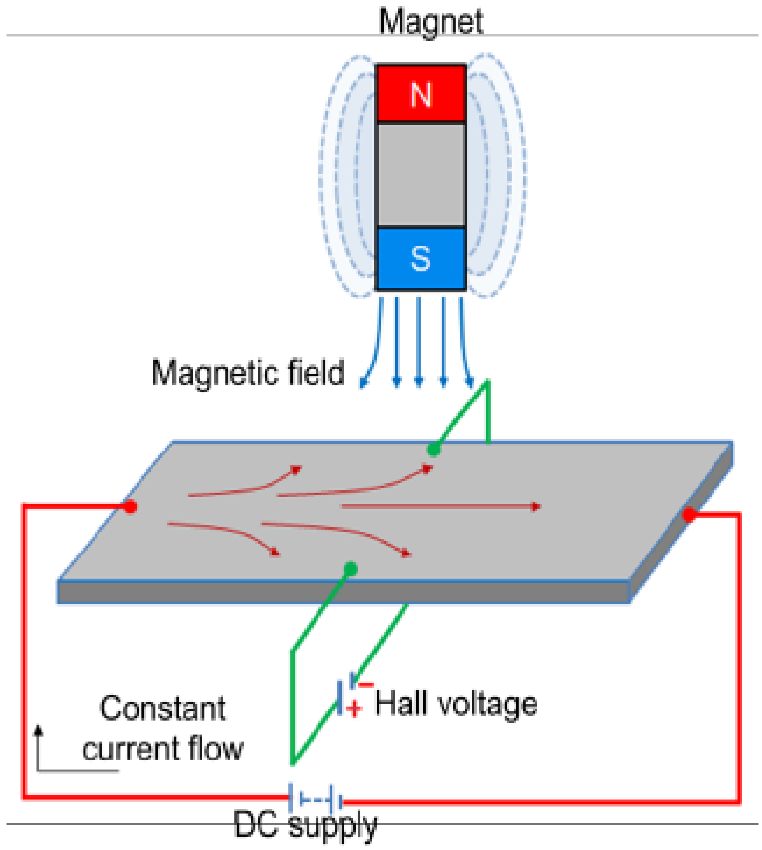


Figure 2.3.8 Principle of working of Hall effect sensor

Figure 2.3.8 shows the principle of working of Hall effect sensor. Hall effect sensors work on the principle that when a beam of charge particles passes through a magnetic field, forces act on the particles and the current beam is deflected from its straight line path. Thus one side of the disc will become negatively charged and the other side will be of positive charge. This charge separation generates a potential difference which is the measure of distance of magnetic field from the disc carrying current.

The typical application of Hall effect sensor is the measurement of fluid level in a container. The container comprises of a float with a permanent magnet attached at its top. An electric circuit with a current carrying disc is mounted in the casing. When the fluid level increases, the magnet will come close to the disc and a potential difference generates. This voltage triggers a switch to stop the fluid to come inside the container.

RTDs are used in the form of thin films, wire wound or coil. They are generally made of metals such as platinum, nickel or nickel-copper alloys. Platinum wire held by a high-temperature glass adhesive in a ceramic tube is used to measure the temperature in a metal furnace. Other applications are:

- Air conditioning and refrigeration servicing
- Food Processing
- Stoves and grills
- Textile production
- Plastics processing
- Petrochemical processing
- Micro electronics
- Air, gas and liquid temperature measurement in pipes and tanks
- Exhaust gas temperature measurement

3. Thermistors

Thermistors follow the principle of decrease in resistance with increasing temperature. The material used in thermistor is generally a semiconductor material such as a sintered metal oxide (mixtures of metal oxides, chromium, cobalt, iron, manganese and nickel) or doped polycrystalline ceramic containing barium titanate ($BaTiO_3$) and other compounds. As the temperature of semiconductor material increases the number of electrons able to move about increases which results in more current in the material and reduced resistance. Thermistors are rugged and small in dimensions. They exhibit nonlinear response characteristics.

Thermistors are available in the form of a bead (pressed disc), probe or chip. Figure 2.5.4 shows the construction of a bead type thermistor. It has a small bead of dimension from 0.5 mm to 5 mm coated with ceramic or glass material. The bead is connected to an electric circuit through two leads. To protect from the environment, the leads are contained in a stainless steel tube.

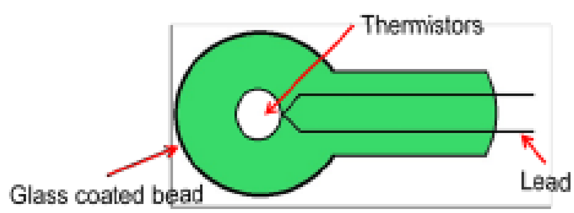


Figure 2.5.4 Schematic of a thermistor

The Oscilloscope: Operation and Applications

1. The Oscilloscope

Oscilloscope Operation (X vs Y mode)

An oscilloscope can be used to measure voltage. It does this by measuring the voltage drop across a resistor and in the process draws a small current. The voltage drop is amplified and used to deflect an electron beam in either the X (horizontal) or Y (vertical) axis using an electric field. The electron beam creates a bright dot on the face of the Cathode Ray Tube (CRT) where it hits the phosphorous. The deflection, due to an applied voltage, can be measured with the aid of the calibrated lines on the graticule.

First we will consider the circuitry that amplifies and conditions the voltage to be measured (the "Amp" block in figure 1).

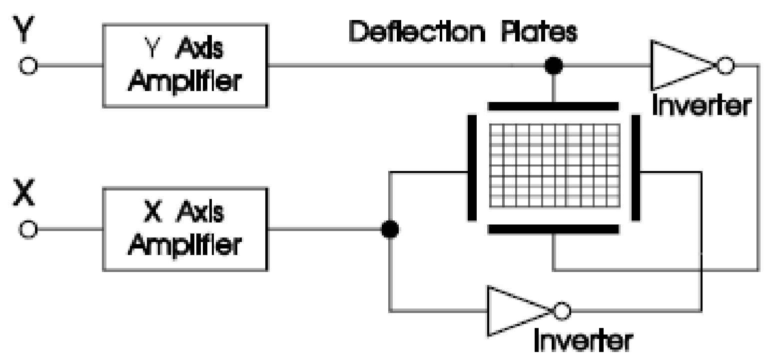


Figure 1. X vs. Y Deflection Block Diagram of the CRT

The deflection of the oscilloscope beam is proportional to the input voltage (after ac or dc coupling). The amount of deflection (Volts/Division) depends upon the setting of the AMPL/DIV control for that channel (see figure 2).

The input signal can be ac or dc coupled. Ac coupling involves adding a series capacitor. This has the effect of blocking (removing) the dc bias and low frequency components of a signal. Dc coupling does not have this problem and therefore allows you to measure voltages right down to 0 Hz. Ac coupling is useful when you are trying to measure a small ac voltage that is "on-top" of a large dc voltage. A typical example is trying to measure the noise of a dc power supply.

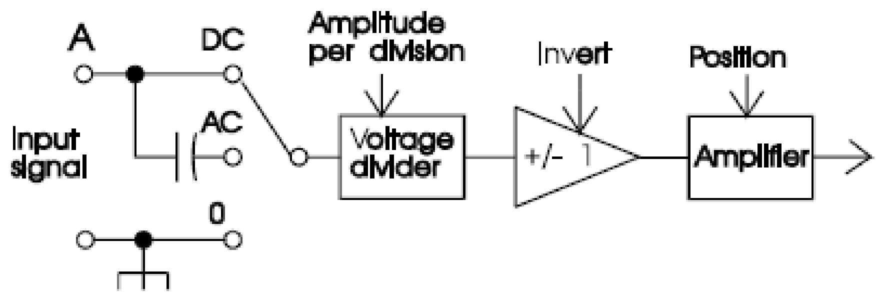


Figure 2. Amplifier Block Diagram

Amplifier Features

AMPL/DIV - This abbreviated name varies but it is generally some short form of amplitude per division. The control is a simple voltage divider (attenuator) which is used to change the sensitivity of the oscilloscope. At a 1 volt/DIV setting, a deflection of one major division on the graticule represents a one volt change at the oscilloscope input.

Calibrated voltage measurements

The small knob within the AMPL/DIV control must be rotated clockwise into its detente position for the amplifiers to be calibrated. Otherwise the voltage/division will be some unknown value greater than what the dial indicates.

INV - There is almost always a control which lets you invert one channel. This can be used along with the ADD function to subtract two voltages. This is necessary because the common input (black lead of the oscilloscope cable) can only be connected to a 0V node. If channel A has V1 + V2 and channel B has voltage V1 then the reading of channel A + (-channel B) = (V1 + V2) + (-V1) = V2

Position - For each axis there is a control which lets you shift the electron beam. With this you can set the zero voltage point to anywhere that is convenient for you.

Oscilloscope Inputs

The input of the oscilloscope can usually be modelled as a resistance and a parallel capacitance (see figure 3). The resistance is usually 1MΩ but it and the capacitance can vary greatly. The total or effective capacitance includes the oscilloscope circuitry (approx. 30 pF), cables (approx. 30 pF/m) and stray capacitance. The resistance will draw current from the circuit while the capacitance will add an RC time constant with its associated time delay, frequency response and distortion of some waveforms.

The common connection (black lead or shield) at the input of the oscilloscope goes to the metal case as the symbol by the input connector shows. Because of this, the common input can only be connected to a 0V point in the circuit. Since the common inputs for both the A and B channels are connected to the case, they are effectively shorted together.

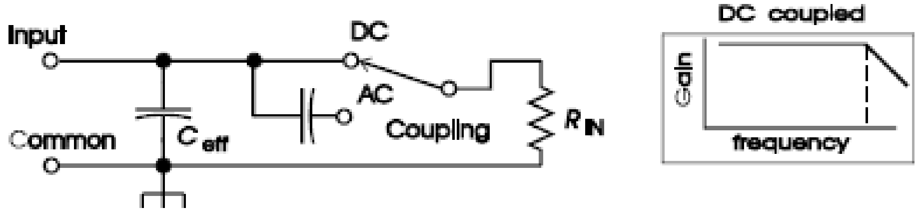


Figure 3. DC-coupled Oscilloscope Input Circuit and Frequency Response

Frequency response is calculated or measured by applying a pure sinusoidal waveform to a circuit. The circuit response is the output voltage divided by the input voltage. This is a complex number that can also be expressed as a magnitude (gain) and phase.

Due to limitations in the amplifiers, the oscilloscope's frequency response is limited. The manufacturer simply lists the half-power point for the oscilloscope without any external effects. Half power is also called the -3dB point. At this point, the voltage has decreased to 70.7% of its maximum. This means that only one-half of the maximum power would be dissipated in a resistive load. Keep in mind that an oscilloscope that is rated at 20 MHz is usually only accurate to 4 MHz for non-sinusoidal waveforms before distortion becomes a problem.

With ac coupling (figure 4), an oscilloscope has another series RC circuit. It acts like a high pass filter (HPF). If you are viewing low frequency signals when ac coupled, not only will you not be able to measure any dc offset, but you will also be removing some low frequency information.

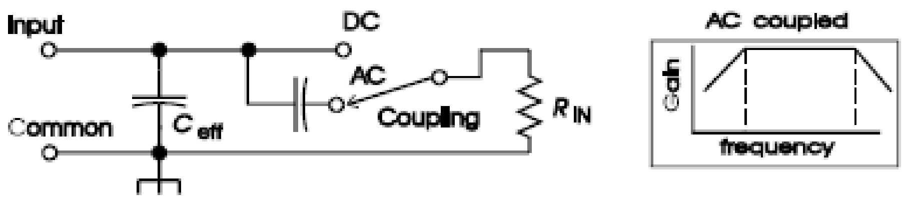


Figure 4. AC-coupled Oscilloscope Input Circuit and Frequency Response

Oscilloscope Operation (Voltage vs Time)
The main function of an oscilloscope is to show voltage vs time. This is done by applying a ramp (or sawtooth) waveform into the X-axis amplifier as shown in figure 5. During the rising edge of the ramp, the electron beam scans across the screen. When the voltage drops back to 0V, the beam is turned off and quickly goes back to its starting point. This is signified by a thick line when the beam is on and a thin one when it is off (blanked).

To obtain a stable picture on the CRT screen, the ramp waveform has to be in phase with the signal that you want to observe. This is done with a triggering circuit. The triggering circuit allows the oscilloscope to draw repeatedly the same waveform over and over by identifying the same point on a repetitive waveform.

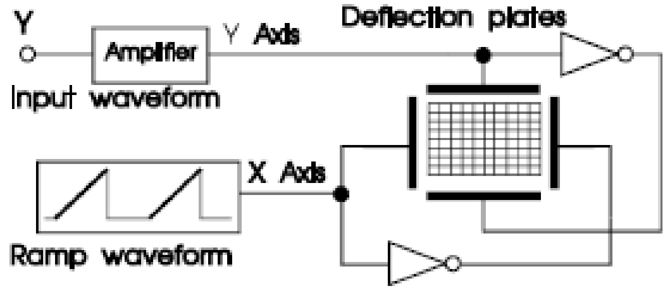


Figure 5. A Ramp-driven X-axis input

The triggering circuit allows you to select a voltage (an analog value) and an edge or slope (positive or negative) for the triggering circuit to compare to the input waveform. When the two are equal, the circuit puts out a pulse. This pulse triggers the ramp waveform generator to do one cycle of its rising and falling edges. Once the ramp has started a cycle of increasing voltage, it can not be retriggered until it has completed the full ramp and returned to 0V. This is illustrated in figure 6 for a single cycle and in figure 7 for multiple cycles.

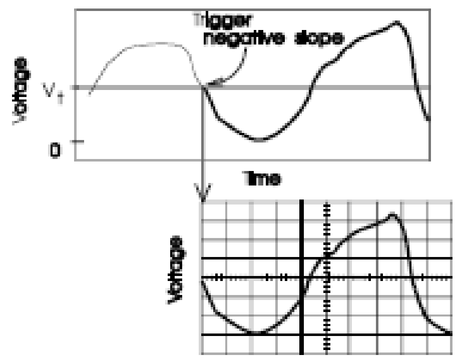


Figure 6. A Triggering Example

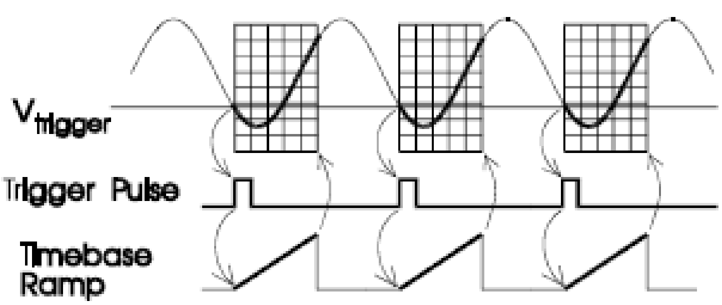


Figure 7. Several Triggering Cycles

Not only do you have control over the starting point of the ramp, but the amount of time that the ramp takes to reach its maximum voltage (the right hand side of the CRT screen) can be adjusted with the timebase control. In essence, you have a “window”. You can move the window to any point on a waveform with the triggering circuit and you can change the size of the window with the timebase.

The time-base control allows you to set the time / division that the beams takes to scan across the screen. Just like the voltage selector, there is a calibration knob in the middle of the control. Unless the vernier (calibration knob) is 'clicked' in to its most clockwise position, the time per division is unknown.

When set to AUTO (automatic) triggering, the oscilloscope will always show a trace. However, when you use a manual triggering mode (DC, AC), many strange things can happen. For example, if the triggering voltage or level is set to +10V and the waveform never exceeds +5V, the triggering circuit will never trigger and the screen will stay blank.

You may think that in a condition of no triggering, you would still have a bright dot on the screen because the electron beam would go to its 'home' or undeflected position. Since the oscilloscope is designed to work with a moving electron beam, a stationary beam can very quickly 'burn' a hole in the phosphorous coating of the screen. To prevent this, there is a ' blanking' circuit which turns off the electron beam. Blanking occurs when there is no triggering or when the electron beam is sweeping from the right edge back to the left side of the screen.

Time measurements are done the same way as voltage measurements. As long as the timebase is calibrated you multiply the number of divisions by the number of seconds per division to get the total time difference. Phase measurements are done by comparing the measured time to the period of the waveform.

Oscilloscope Two Channel Operation
You can view two voltage waveforms at once by using two Y-axis (vertical) input channels. The individual channels are sometimes labelled as '1' and '2' or as 'A' and 'B'. Since there is only one electron beam, you have to share its drawing time between both waveforms. This may be accomplished using either the **chop** or **alternate** modes.

When in the chop mode (figure 8), the oscilloscope displays a little bit of channel A, then a little bit of B, then A, then B ....during a single sweep of the electron beam. If you increase the timebase to about 1µs/division, you can start to see the individual pieces as it chops between one channel and the other channel.

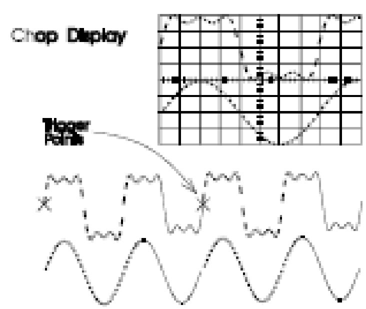


Figure 8. Chop Mode

In the alternate mode (figure 9), the oscilloscope will sweep the electron beam twice across the screen. The first time it will draw the signal from channel A and the next time from channel B. At very low timebase settings, you can see it draw one channel and then the other in successive passes.
Note: When you use the alternate function, the two waveforms that you see are from different points in time and the triggering circuit has to trigger twice.

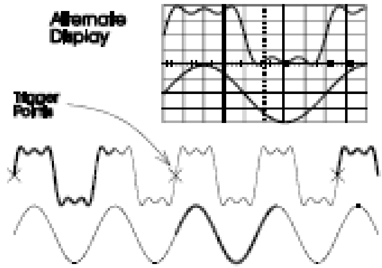


Figure 9. Alternate Mode

The reason that you can see a non-flickering image on the screen is because the phosphorous coating on the CRT has persistence. In essence, the phosphorous acts like a low pass filter and averages several images that are drawn on the screen.

By viewing two signals at a time, you can measure relative time differences. By combining a voltage and phase measurement (relative to the appropriate reference), you can measure a phasor value.

With a two channel oscilloscope, you have the ability to trigger on each waveform and electronically switch (chop or alt) between them as well. A block diagram of a oscilloscope has now become as shown in figure 10.

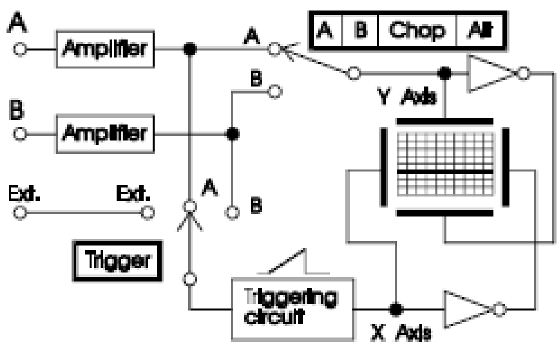


Figure 10. A Simple Oscilloscope Block Diagram

Some oscilloscopes offer a way to alternately trigger as depicted in figure 11. When combined with alternate displaying, you can stably display two waveforms of any frequency by alternately showing each channel and triggering on the channel that is being drawn. This way, the oscilloscope is acting like a two beam scope with both waveforms triggering at the same voltage and slope. However, there is no way to know what the relationship is between one waveform and the other when using alternate triggering.

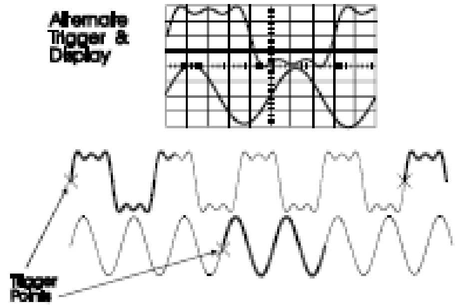


Figure 11. Stable Triggering of Two Different Frequencies

If you have two waveforms that do not have the same frequency, it is still possible to show them as two stable waveforms on a normal oscilloscope. In figure 11, you will notice that if the triggering occurs at the 'X', both waveforms are in phase (ie. at the same phase each time the timebase triggers). The condition for a stable display is not that two waveforms have to be of exactly the same frequency, but that when they are triggered, they have to be in phase.
Or $n \cdot f1 = m \cdot f2$ where n, m are integers. That is not necessary, but it is sufficient. There are many other ways to achieve a stable trace when you consider that the trigger circuit will wait for the next triggering point. There is also a control on some oscilloscopes, called 'hold off', which allows you to add a delay between the end of the trace being drawn and the time when the triggering circuit starts to look for the next triggering point. That can be used to stabilize the display under some circumstances.

Remember that all of this applies only for repetitive waveforms that are properly triggered. If the triggering is not stable, or the waveform is not repetitive, you will see a constantly moving image or several images offset and superimposed.

A slightly more complicated block diagram of an oscilloscope, with the typical functions found in the laboratory, is illustrated in figure 12.

Accuracy
There are many factors affecting the accuracy of oscilloscope measurements.
There are errors due to the input channel voltage divider, timebase control, the use of magnifiers, the accuracy to which the CRT deflection can be read, beam thickness, temperature etc. The voltage divider error will be the same for all readings that are done on the same timebase and voltage range, but may be different each time the range is changed. Measurements over only two divisions can incur two to three times the error of those made over the centre eight divisions.

If the phase angle is used in a trigonometric function, this error can be multiplied by the slope of the function. Consider that the tangent of a 1% phase error entered at 85 degrees is much worse (20%) than the same 1% error on a sine function (0.2%) at the same angle. To get a feel for this look at the Taylor expansion of the trigonometric functions.

It is wise to consult the user manual for a particular instrument's accuracy specifications.

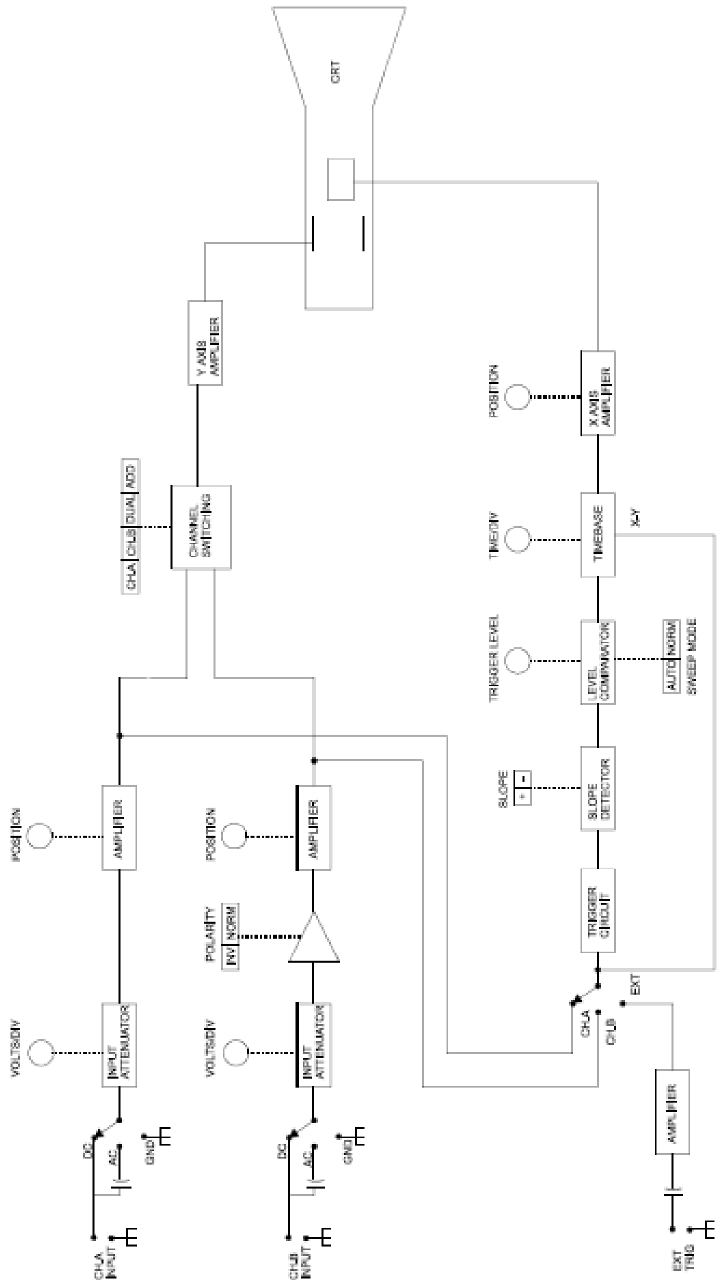


Figure 12. Oscilloscope Block Diagram

2. Iwatsu Model SS-5702 Oscilloscope

Accuracy

The Iwatsu SS-5702 oscilloscope is used in this laboratory. All measurements on the graticule should be made over as many divisions as possible. For simplicity, assume the Iwatsu SS-5702 oscilloscope's error is ±5% for a measurement on either the vertical or horizontal scales over eight divisions.

Front Panel Controls

The front panel controls, shown in figure 13, will be described in the remainder of this section.

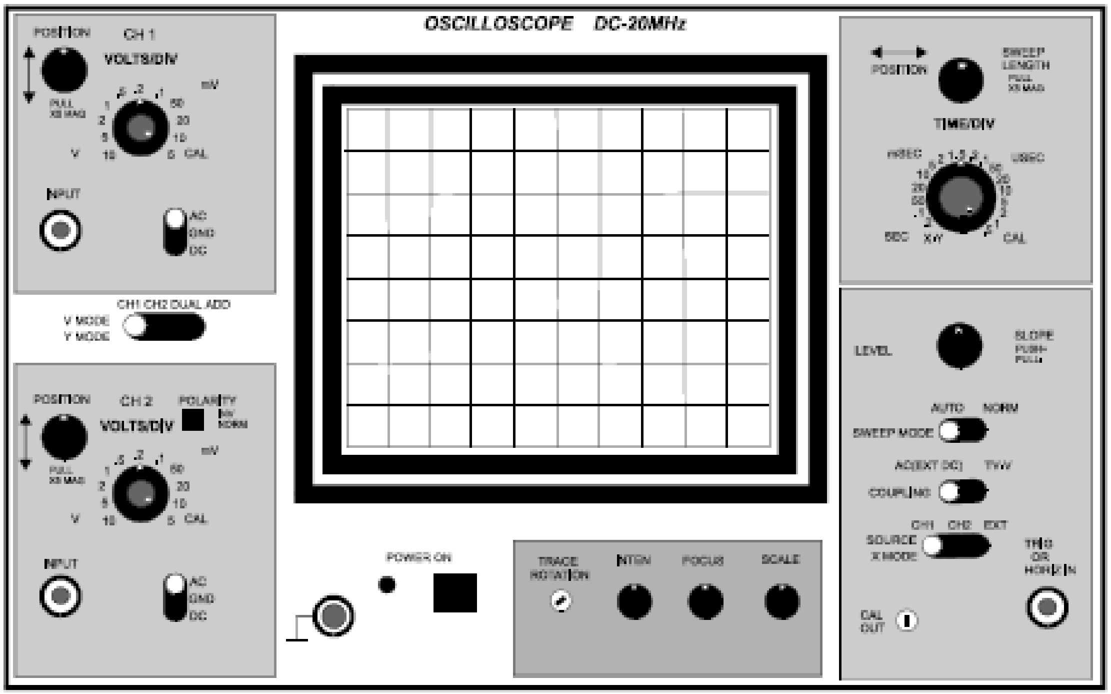


Figure 13. The Iwastu SS-5702 Oscilloscope

Front Panel Controls

The power, trace rotation, intensity, focus, and scale illumination controls are located at the bottom centre.
The vertical controls and input selection are on the left side.
The horizontal controls and horizontal input selection are on the right side.
The triggering selection section is at the bottom right.

The power on/off, trace rotation, intensity, focus and scale illumination

Trace rotation has to be checked with just a straight line across the screen of the oscilloscope.

The intensity should be adjusted to a mid point (or more clockwise).
The focus of the beam can be done while observing the straight line display.
The scale illumination can be adjusted to the operator's preference.

Vertical Inputs
Vertical channels are used for measuring voltage.
The beam is deflected vertically as a result of the signal being applied to the vertical input of the channel (CH).

CH-1 and **CH-2** are the labels for the two vertical inputs.

Each channel has a position control, a range selection switch, a pull "x5" knob, a coupling selector switch (AC/GND/DC), and an input connector.

Each channel Range switch (VOLTS/DIV) has a smaller knob in the middle of the Range Selector Switch. And there is an arrow showing that the Range Selector Switch is in the **"CAL"** position when rotated fully clockwise.

In the centre of the two channel sections is the channel selection switch. You can choose to have CH-1, CH-2, both (DUAL), or ADD.

In addition, **CH-2** has a **"Polarity"** switch. You push the polarity switch in to **"INVERT"** the polarity of the signal being displayed. The **"NORM"** or out position is the normal position of the polarity switch.

Horizontal Inputs
The **"EXT"** (HORIZONTAL IN) can be supplied a voltage directly via the connector at the bottom right of the panel, or the Horizontal can be driven by an internal timebase circuit which generates the voltage.

Timebase
In the timebase mode, the horizontal signal is from an internal source which changes **linearly with respect to time**. Hence, the beam is deflected to give us a calibration of time for the horizontal scale.

The position control, the pull "x5" magnifier switch, the time range selection switch, and the range "CAL" knob all affect the X-axis of the display.

Trigger Sources
The TRIGGER SOURCE may be selected from one of three sources, **CH-1, CH-2, or EXT.**
Look at the bottom right of the control panel.

Calibration Source
The Calibration Source is an internal source, available on the oscilloscope.
Look at the bottom right side of the front panel. The output is labelled 0.3 V.
This is a 1000 Hz, square-wave (the 50 % duty cycle is not accurate).

3. Oscilloscope Applications

Voltage and Time Measurements
Note: **The oscilloscope measures divisions of deflection not voltage or time**. From the divisions of deflection you can calculate the time or voltage.

Differential Measurements
An important application of the oscilloscope is differential measurements. Such measurements are necessary because both vertical channels have one terminal connected to the chassis common (ie single ended). To measure a floating (off ground) voltage you have to use the "invert and add" feature of the oscilloscope. For example, in figure 14, to measure V1:

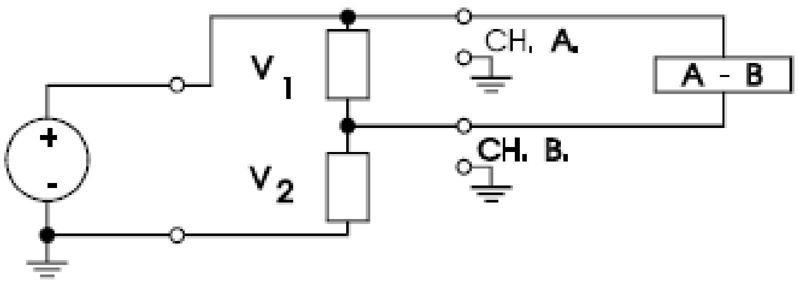


Figure 14. Differential Measurement Example

Channel A measures (V1 + V2) relative to ground while channel B measures V2 relative to ground. By pushing the invert button you negate the voltage displayed on channel B. Then you can add the channels together with the "ADD" display mode. The waveform now displayed is (V1 + V2) + (- V2) = V1.

Bandwidth (-3 dB) Measurement
This measurement is easily done by first finding the maximum gain (max. VOUT / VIN at frequency ω_o) and **adjusting the oscilloscope** so that the sinewave fills seven divisions peak to peak. A -3dB point can be found by increasing and/or decreasing the frequency until the gain is reduced to $\sqrt{2}$. **If the input voltage has remained constant** this will occur when the output voltage is five divisions peak to peak.

$$\frac{V_{out}}{V_{out\text{-}max}} = \frac{1}{\sqrt{2}} = 0.707 \approx \frac{5}{7} \textit{ divisions}$$

The frequency is then simply read with a frequency counter or the oscilloscope. **Not by reading the dial of the signal generator**.

Risetime
The risetime indicates how quickly a circuit responds. The risetime is the time it takes a waveform to go from 10% of the voltage range to 90% of the voltage range. This is in response to a square wave and the output voltage must settle to a steady-state voltage (0% and 100%). Most oscilloscopes have dotted lines on the graticule marking the 10% and 90% points to aid in this measurement. Usually these dotted lines assume that 0% is the lowest line of the graticule and 100% is the highest line. The measurement, as shown in figure 15, also includes the risetime of the oscilloscope and the squarewave source.

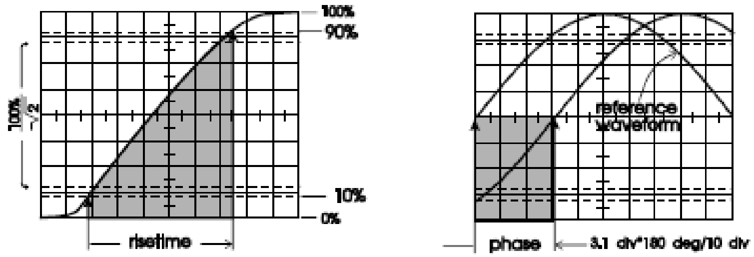


Figure.15 A Risetime and Phase Measurement

Phase

Phase is most accurately measured when the waveform is as large as possible and the difference is measured **at the zero crossings**. Typically the timebase is uncalibrated so that a 180 degree section of the waveform is expanded to the full 10 divisions of the graticule. Then the sign of the phase can be determined by observing more than one period of both waveforms. Both waveforms must be symmetrical about the centre line of the graticule. The angle is determined by: phase = # of divisions * 180 degrees / 10 divisions.

Module 2: Sensors and signal processing
Lecture 1
Sensors and transducers

Measurement is an important subsystem of a mechatronics system. Its main function is to collect the information on system status and to feed it to the micro-processor(s) for controlling the whole system.

Measurement system comprises of sensors, transducers and signal processing devices. Today a wide variety of these elements and devices are available in the market. For a mechatronics system designer it is quite difficult to choose suitable sensors/transducers for the desired application(s). It is therefore essential to learn the principle of working of commonly used sensors/transducers. A detailed consideration of the full range of measurement technologies is, however, out of the scope of this course. Readers are advised to refer "Sensors for mechatronics" by Paul P.L. Regtien, Elsevier, 2012 [2] for more information.

Sensors in manufacturing are basically employed to automatically carry out the production operations as well as process monitoring activities. Sensor technology has the following important advantages in transforming a conventional manufacturing unit into a modern one.

1. Sensors alarm the system operators about the failure of any of the sub units of manufacturing system. It helps operators to reduce the downtime of complete manufacturing system by carrying out the preventative measures.
2. Reduces requirement of skilled and experienced labors.
3. Ultra-precision in product quality can be achieved.

Sensor

It is defined as an element which produces signal relating to the quantity being measured [1]. According to the Instrument Society of America, sensor can be defined as "*A device which provides a usable output in response to a specified measurand.*" Here, the output is usually an 'electrical quantity' and measurand is a 'physical quantity, property or condition which is to be measured'. Thus in the case of, say, a variable inductance displacement element, the quantity being measured is displacement and the sensor transforms an input of displacement into a change in inductance.

The Oscilloscope: Operation and Applications

1. The Oscilloscope

Oscilloscope Operation (X vs Y mode)
An oscilloscope can be used to measure voltage. It does this by measuring the voltage drop across a resistor and in the process draws a small current. The voltage drop is amplified and used to deflect an electron beam in either the X (horizontal) or Y (vertical) axis using an electric field. The electron beam creates a bright dot on the face of the Cathode Ray Tube (CRT) where it hits the phosphorous. The deflection, due to an applied voltage, can be measured with the aid of the calibrated lines on the graticule.
First we will consider the circuitry that amplifies and conditions the voltage to be measured (the "Amp" block in figure 1).

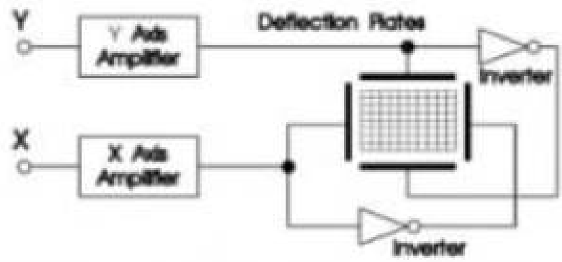


Figure 1. X vs. Y Deflection Block Diagram of the CRT

The deflection of the oscilloscope beam is proportional to the input voltage (after ac or dc coupling). The amount of deflection (Volts/Division) depends upon the setting of the AMPL/DIV control for that channel (see figure 2).

The input signal can be ac or dc coupled. Ac coupling involves adding a series capacitor. This has the effect of blocking (removing) the dc bias and low frequency components of a signal. Dc coupling does not have this problem and therefore allows you to measure voltages right down to 0 Hz. Ac coupling is useful when you are trying to measure a small ac voltage that is "on-top" of a large dc voltage. A typical example is trying to measure the noise of a dc power supply.

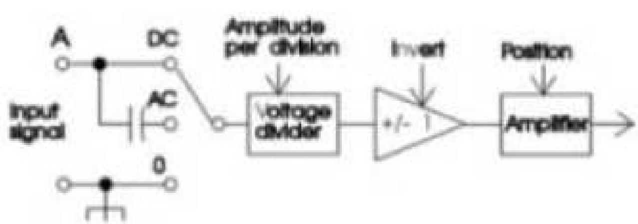


Figure 2. Amplifier Block Diagram

Amplifier Features
AMPL/DIV - This abbreviated name varies but it is generally some short form of amplitude per division. The control is a simple voltage divider (attenuator) which is used to change the sensitivity of the oscilloscope. At a 1 volt/DIV setting, a deflection of one major division on the graticule represents a one volt change at the oscilloscope input.

Calibrated voltage measurements
The small knob within the AMPL/DIV control must be rotated clockwise into its detente

as "*A device which provides a usable output in response to a specified measurand.*" Here, the output is usually an 'electrical quantity' and measurand is a 'physical quantity, property or condition which is to be measured'. Thus in the case of, say, a variable inductance displacement element, the quantity being measured is displacement and the sensor transforms an input of displacement into a change in inductance.

Transducer

It is defined as an element when subjected to some physical change experiences a related change [1] or an element which converts a specified measurand into a usable output by using a transduction principle.

It can also be defined as a device that converts a signal from one form of energy to another form.

A wire of Constantan alloy (copper-nickel 55-45% alloy) can be called as a sensor because variation in mechanical displacement (tension or compression) can be sensed as change in electric resistance. This wire becomes a transducer with appropriate electrodes and input-output mechanism attached to it. Thus we can say that 'sensors are transducers'.

Sensor/transducers specifications

Transducers or measurement systems are not perfect systems. Mechatronics design engineer must know the capability and shortcoming of a transducer or measurement system to properly assess its performance. There are a number of performance related parameters of a transducer or measurement system. These parameters are called as sensor specifications.

Sensor specifications inform the user to the about deviations from the ideal behavior of the sensors. Following are the various specifications of a sensor/transducer system.

1. Range

The range of a sensor indicates the limits between which the input can vary. For example, a thermocouple for the measurement of temperature might have a range of 25-225 °C.

2. Span

The span is difference between the maximum and minimum values of the input. Thus, the above-mentioned thermocouple will have a span of 200 °C.

3. Error

Error is the difference between the result of the measurement and the true value of the quantity being measured. A sensor might give a displacement reading of 29.8 mm, when the actual displacement had been 30 mm, then the error is –0.2 mm.

3. Oscilloscope Applications

Voltage and Time Measurements

Note: The **oscilloscope measures divisions of deflection not voltage or time**. From the divisions of deflection you can calculate the time or voltage.

Differential Measurements

An important application of the oscilloscope is differential measurements. Such measurements are necessary because both vertical channels have one terminal connected to the chassis common (ie single ended). To measure a floating (off ground) voltage you have to use the "invert and add" feature of the oscilloscope. For example, in figure 14, to measure V1:

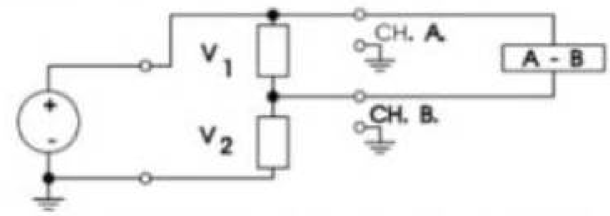


Figure 14. Differential Measurement Example

Channel A measures (V1 + V2) relative to ground while channel B measures V2 relative to ground. By pushing the invert button you negate the voltage displayed on channel B. Then you can add the channels together with the "ADD" display mode. The waveform now displayed is (V1 + V2) + (- V2) = V1.

Bandwidth (-3 dB) Measurement

This measurement is easily done by first finding the maximum gain (max. VOUT / VIN at frequency ω_o) and **adjusting the oscilloscope** so that the sinewave fills seven divisions peak to peak. A -3dB point can be found by increasing and/or decreasing the frequency until the gain is reduced to $\sqrt{2}$. **If the input voltage has remained constant** this will occur when the output voltage is five divisions peak to peak.

$$\frac{V_{out}}{V_{out\text{-}max}} = \frac{1}{\sqrt{2}} = 0.707 = \frac{5}{7}\ \textit{divisions}$$

The frequency is then simply read with a frequency counter or the oscilloscope. **Not by reading the dial of the signal generator**.

Risetime

The risetime indicates how quickly a circuit responds. The risetime is the time it takes a waveform to go from 10% of the voltage range to 90% of the voltage range. This is in response to a square wave and the output voltage must settle to a steady-state voltage (0% and 100%). Most oscilloscopes have dotted lines on the graticule marking the 10% and 90% points to aid in this measurement. Usually these dotted lines assume that 0% is the lowest line of the graticule and 100% is the highest line. The measurement, as shown in figure 15, also includes the risetime of the oscilloscope and the squarewave source.

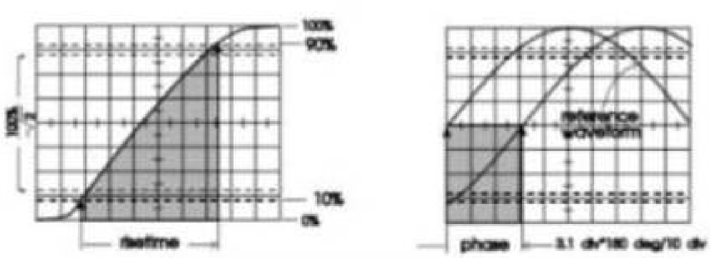


Figure.15 A Risetime and Phase Measurement

Phase

Phase is most accurately measured when the waveform is as large as possible and the difference is measured **at the zero crossings**. Typically the timebase is uncalibrated so that a 180 degree section of the waveform is expanded to the full 10 divisions of the graticule. Then the sign of the phase can be determined by observing more than one period of both waveforms. Both waveforms must be symmetrical about the centre line of the graticule. The angle is determined by: phase = # of divisions * 180 degrees / 10 divisions.

It is wise to consult the user manual for a particular instrument's accuracy specifications.

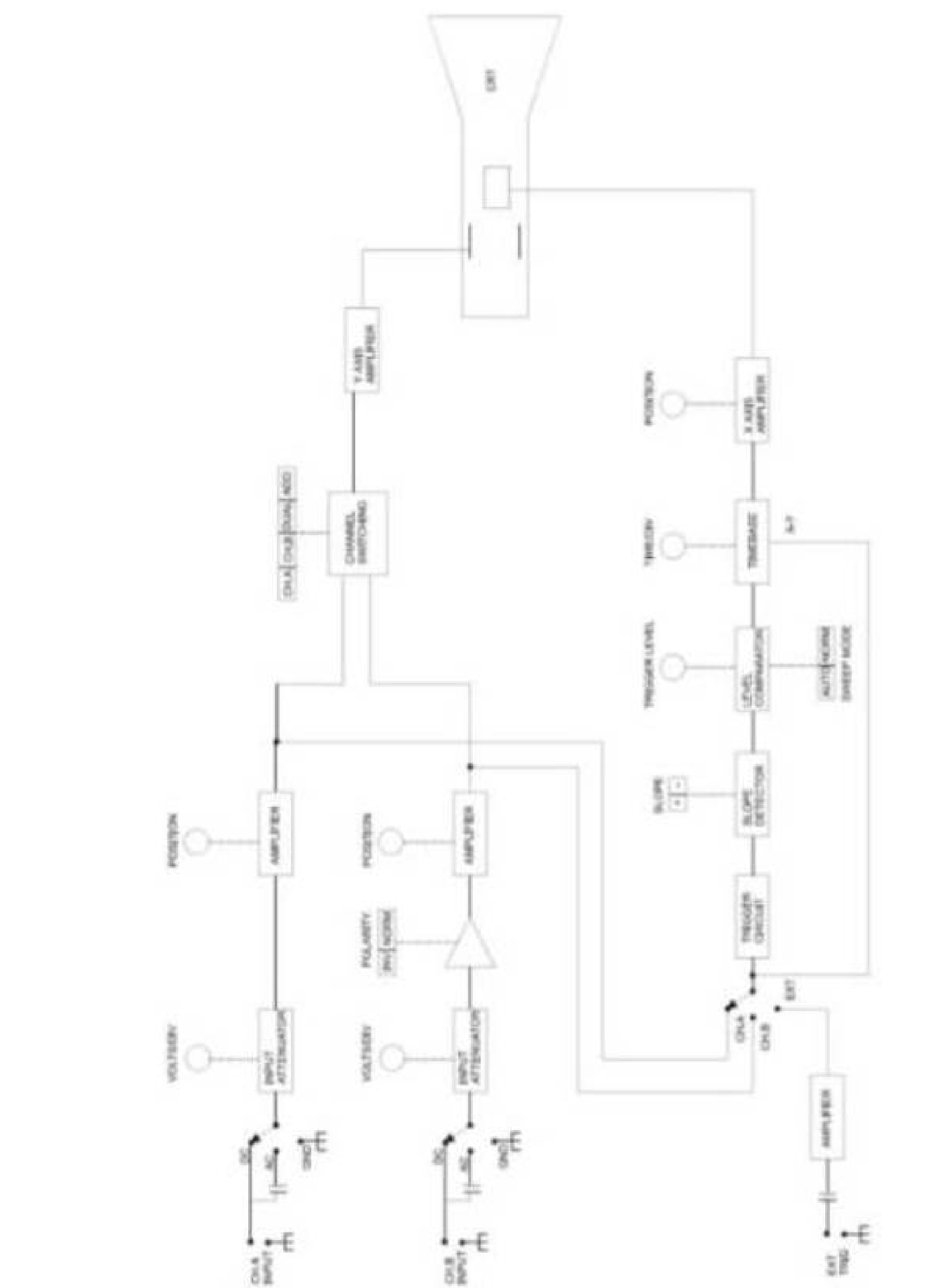


Figure 12. Oscilloscope Block Diagram

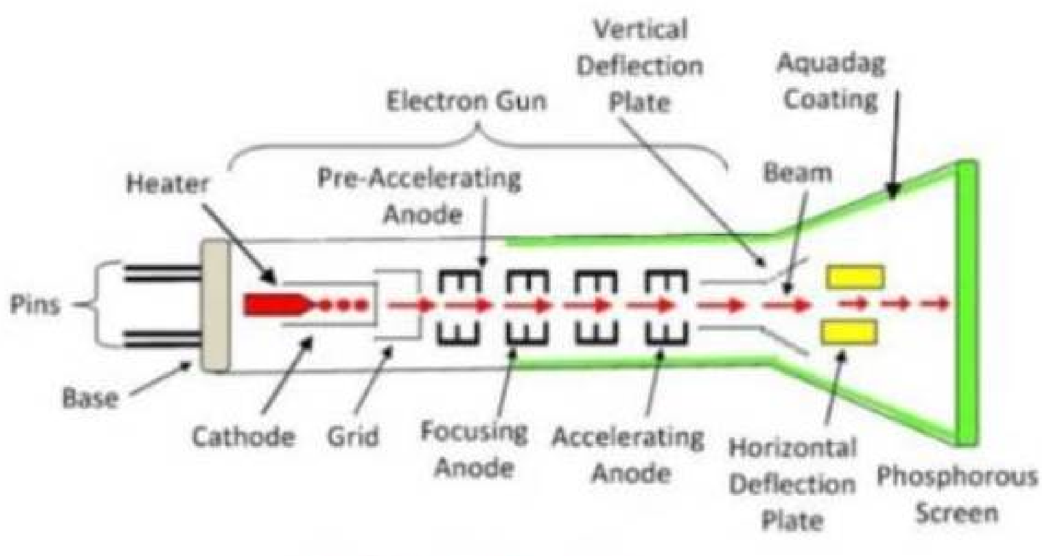
2. Iwatsu Model SS-5702 Oscilloscope

Accuracy

The Iwatsu SS-5702 oscilloscope is used in this laboratory. All measurements on the graticule should be made over as many divisions as possible. For simplicity, assume the Iwatsu SS-5702 oscilloscope's error is ±5% for a measurement on either the vertical or horizontal scales over eight divisions.

Working of CRT

The working of CRT depends on the movement of electrons beams. The electron guns generate sharply focused electrons which are accelerated at high voltage. This high-velocity electron beam when strikes on the fluorescent screen creates luminous spot



Cathode Ray Tube

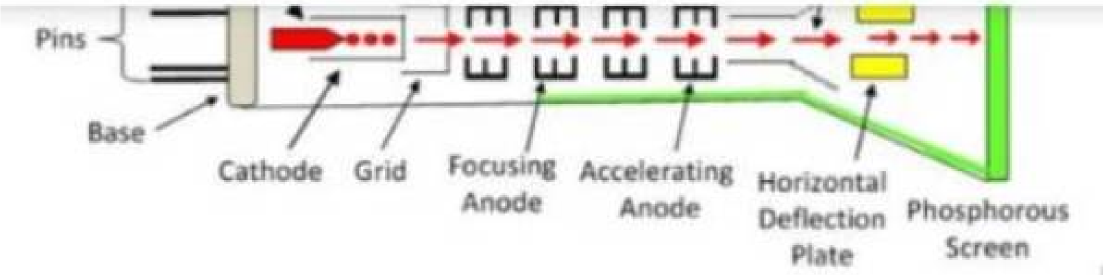
The working parts of a CRT are enclosed in a vacuum glass envelope so that the emitted electron can easily move freely from one end of the tube to the other.

Construction of CRT

The Electrons Gun Assembly, Deflection Plate Assembly, Fluorescent Screen, Glass Envelope, Base are the important parts of the CRT. The electron gun emits the electron beam, and through deflecting plates, it is strikes on the phosphorous screen. The detail explanation of their parts is explained below.

Electrons Gun Assembly

The electron gun is the source of the electron beams. The electron gun has a heater, cathode, grid, pre-accelerating anode, focusing anode and accelerating anode. The electrons are emitted from



Cathode Ray Tube

After exiting from the electron gun, the beam passes through the pairs of electrostatic deflection plate. These plates deflected the beams when the voltage applied across it. The one pair of plate moves the beam upward and the second pair of plate moves the beam from one side to another. The horizontal and vertical movement of the electron are independent of each other, and hence the electron beam positioned anywhere on the screen.

The working parts of a CRT are enclosed in a vacuum glass envelope so that the emitted electron can easily move freely from one end of the tube to the other.

The deflection plate produces the uniform electrostatic field only in the one direction. The electron beam entering into the deflection plates will accelerate only in the one direction, and hence electrons will not move in the other directions.

Screen For CRT

The front of the CRT is called the face plate. The face plate of the CRT is made up of entirely fibre optics which has special characteristics. The internal surface of the faceplate is coated with the phosphor. The phosphorous converts the electrical energy into light energy. The energy level of the phosphorous crystal raises when the electron beams strike on it. This phenomenon is called cathodoluminescence.

the electron beam, and through deflecting plates, it is strikes on the phosphorous screen. The detail explanation of their parts is explained below.

Electrons Gun Assembly

The electron gun is the source of the electron beams. The electron gun has a heater, cathode, grid, pre-accelerating anode, focusing anode and accelerating anode. The electrons are emitted from the highly emitted cathode. The cathode is cylindrical in shape, and at the end of it, the layer of strontium and barium oxide is deposited which emit the high emission of electrons at the end of the tube.

GANDHI ACADEMY OF TECHNOLOGY AND ENGINEERING
(Approved By AICTE New Delhi & Affiliated to BPUT, Rourkela, Odisha)

ENVIRONMENTAL STUDIES

PREPARED BY

Er. SUDHIR KUMAR DAS
DEPT. of Electrical Engineering

Th5. ENVIRONMENTAL STUDIES

(Common to all Branches)

Name of the Course: Diploma in Electrical Engineering			
Course code:		Semester	3rd
Total Period:	60	Examination :	3 hrs
Theory periods:	4P / week	Internal Assessment:	20
Maximum marks:	100	End Semester Examination ::	80

A. RATIONALE:

Due to various aspects of human developments including the demand of different kinds of technological innovations, most people have been forgetting that, the Environment in which they are living is to be maintained under various living standards for the preservation of better health. The degradation of environment due to industrial growth is very much alarming due to environmental pollution beyond permissible limits in respect of air, water industrial waste, noise etc. Therefore, the subject of Environmental Studies to be learnt by every student in order to take care of the environmental aspect in each and every activity in the best possible manner.

B. OBJECTIVE:

After completion of study of environmental studies, the student will be able to:

1. Gather adequate knowledge of different pollutants, their sources and shall be aware of solid waste management systems and hazardous waste and their effects.
2. Develop awareness towards preservation of environment.

C. Topic wise distribution of periods:

Sl. No.	Topics	Period
1	The Multidisciplinary nature of environmental studies	04
2	Natural Resources	10
3	Systems	08
4	Biodiversity and it's Conservation	08
5	Environmental Pollution	12
6	Social issues and the Environment	10
7	Human population and the environment	08
	Total:	**60**

D. COURSE CONTENTS

1. The Multidisciplinary nature of environmental studies:

1.1 Definition, scope and importance.

1.2 Need for public awareness.

2. Natural Resources:

Renewable and non renewable resources:

2.1 Natural resources and associated problems.
- 2.1.1. Forest resources: Use and over-exploitation, deforestation, case studies, Timber extraction mining, dams and their effects on forests and tribal people.
- 2.1.2. Water resources: Use and over-utilization of surface and ground water, floods, drought, conflicts over water, dam's benefits and problems.
- 2.1.3. Mineral Resources: Use and exploitation, environmental effects of extracting and using mineral resources.
- 2.1.4. Food Resources: World food problems, changes caused by agriculture and over grazing, effects of modern agriculture, fertilizers- pesticides problems, water logging, salinity, .
- 2.1.5. Energy Resources: Growing energy need, renewable and non-renewable energy sources, use of alternate energy sources, case studies.
- 2.1.6. Land Resources: Land as a resource, land degradation, man induces landslides, soil erosion, and desertification.

2.2 Role of individual in conservation of natural resources.
2.3 Equitable use of resources for sustainable life styles.

3. Systems:

3.1. Concept of an eco system.
3.2. Structure and function of an eco system.
3.3. Producers, consumers, decomposers.
3.4. Energy flow in the eco systems.
3.5. Ecological succession.
3.6. Food chains, food webs and ecological pyramids.
3.7. Introduction, types, characteristic features, structure and function of the following eco system:
3.8. Forest ecosystem:
3.9. Aquatic eco systems (ponds, streams, lakes, rivers, oceans,

estuaries).

4. **Biodiversity and it's Conservation:**

4.1. Introduction-Definition: genetics, species and ecosystem diversity.
4.2. Biogeographically classification of India.
4.3. Value of biodiversity: consumptive use, productive use, social ethical, aesthetic and optin values.
4.4. Biodiversity at global, national and local level.
4.5. Threats to biodiversity: Habitats loss, poaching of wild life, man wildlife conflicts.

5. **Environmental Pollution:**

5.1. Definition Causes, effects and control measures of:

5.1.1 Air pollution.
5.1.2 Water pollution.
5.1.3 Soil pollution
5.1.4 Marine pollution
5.1.5 Noise pollution.
5.1.6 Thermal pollution
5.1.7 Nuclear hazards.

5.2. Solid waste Management: Causes, effects and control measures of urban and industrial wastes.
5.3. Role of an individual in prevention of pollution.
5.4. Disaster management: Floods, earth quake, cyclone and landslides.

6. **Social issues and the Environment:**

6.1. Form unsustainable to sustainable development.
6.2. Urban problems related to energy.
6.3. Water conservation, rain water harvesting, water shed management.
6.4. Resettlement and rehabilitation of people; its problems and concern.
6.5. Environmental ethics: issue and possible solutions.
6.6. Climate change, global warming, acid rain, ozone layer depletion, nuclear accidents and holocaust, case studies.
6.7. Air (prevention and control of pollution) Act.
6.8. Water (prevention and control of pollution) Act.
6.9. Public awareness.

7. **Human population and the environment:**

7.1. Population growth and variation among nations.
7.2. Population explosion- family welfare program.
7.3. Environment and humanhealth.
7.4. Human rights.
7.5. Value education
7.6. Role of information technology in environment and human health.

The Multidisciplinary nature of environmental studies

UNIT -1

Definition, Scope and Importance, Need for Public Awareness.

INTRODUCTION:-

The word environment is derived from the French word **‘environner’** which means to **‘Encircle or surround’**.

- Thus our environment can be defined **as “the Social, Cultural and Physical conditions that surround, affect and influence the survival, growth and development of people, animals and plants”**
- This broad definition includes the natural world and the technological environment as well as the cultural and social contexts that shape human lives.
- It includes all factors (living and nonliving) that affect an individual organism or population at any point in the life cycle; set of circumstances surrounding a particular occurrence and allthe things that surrounds us.

SEGMENTS OF ENVIRONMENT

Environment consists of four segments.

1) Atmosphere- Blanket of gases surrounding the earth.
2) Hydrosphere- Various water bodies present on the earth.
3) Lithosphere- Contains various types of soils and rocks on the earth.
4) Biosphere- Composed of all living organisms and their interactions with theenvironment.

MULTIDISCIPLINARY NATURE OF ENVIRONMENTALSTUDIES

1) The Environment studies is a multi-disciplinary science because it comprises various branches of studies like chemistry, physics, medical science, life science, agriculture, public health, sanitary engineering etc.
2) It is the science of physical phenomena in the environment. It studies about the sources, reactions, transport, effect and fate of physical and biological species in the air, water, soil and the effect of from human activity upon these.
3) As the environment is complex and actually made up of many different environments like natural, constructed and cultural environments, environmental studies is inter disciplinary in nature including the study of biology, geology, politics, policy studies, law, religion engineering, chemistry and economics to understand the humanity’seffects on the natural world.
4) This subject educates the students to appreciate the complexity of environmental issues and citizens and experts in many fields. By studying environmental science, students may develop a breadth of the inter disciplinary and methodological knowledge in the environmental fields that enables them to facilitate the definition and solution of environmental problems.

SCOPE OF ENVIRONMENTALSTUDIES

Environmental studies as a subject has a wide scope. It includes a large number of areas and aspects, which may be summarized as follows:

- Natural resources- their conservation and management
- Ecology and Biodiversity
- Environmental pollution and control

- Human population and environment
- Social issues in relation to development and environment

These are the basic aspects of environmental studies which have a direct relevance to every section of society. Several career options have emerged in these fields that are broadly categorized as:

1) **Research and development in environment:**

Skilled environmental scientists have an important role to play in examining various environmental problems in a scientific manner and carry out R&D activities for developing cleaner technologies and promoting sustainable development.

2) **Green advocacy:**

With increasing emphasis on implementing various Acts and Laws related to environment, need for environmental lawyers has emerged, who should be able to plead the cases related to water, air, forest, wildlife, pollution and control etc.

3) **Green marketing:**

While ensuring the quality of products with ISO mark, now there is an increasing emphasis on marketing goods that are environment friendly. Such products have Eco mark or ISO 14000 certification. Environmental auditors and environmental managers would be in great demand in the coming years.

4) **Green media:**

Environmental awareness can be spread amongst masses through mass media like television, radio, newspaper, magazine, hoardings, advertisements etc., for which environmentally educated persons are required.

5) **Environmental consultancy:**

Many non-government organizations, industries and government bodies are engaging environmental consultants for systematically studying and tackling environment related problems.

❖ IMPORTANCE OF ENVIRONMENTALSTUDIES

- The importance of environmental studies is that, the current trend of environmental degradation can be reversed if people of educated communities are organized, empowered and experts are involved in sustainable development.
- Environmental factors greatly influence every organism and their activities.
- At present a great number of environmental issues, have grown in size and complexity day by day, threatening the survival of mankind on earth. These issues are studied besides giving effective suggestions in the environment studies.
- The environment studies enlighten, about the importance of protection and conservation of our natural resources, indiscriminate release of pollution into the environment etc.

Environment studies have become significant for the following reasons:

❖ Environment Issues being of International Importance:

It has been well recognized that environment issues like global warming, ozone depletion, acid rain, marine pollution and loss of biodiversity are not merely national issues but are global issues and hence must be tackled with international efforts and cooperation.

❖ **Problems Cropped in the Wake of Development:**

Development, in its wake gave birth to Urbanization, Industrial Growth, Transportation Systems, Agriculture and Housing etc. However, it has become phased out in the developed world. The North, to cleanse their own environment has fact fully, managed to move ‘dirty’ factories to South. When the West developed, it did so perhaps in ignorance of the environmental impact of its activities. Evidently such a path is neither practicable nor desirable, even if developing world follows that.

❖ **Explosively Increase in Pollution:**

World census reflects that one in every seven persons in this plant lives in India. Evidently with 16 per cent of the world's population and only 2.4 per cent of its land area, there is a heavy pressure on the natural resources including land. Agricultural experts have recognized soils health problems like deficiency of micronutrients and organic matter, soil salinity and damage of soil structure.

❖ **Need for an Alternative Solution:** It is essential, especially for developing countries to find alternative paths to an alternative goal.We need a goal as under:

- A goal, which ultimately is the true goal of development an environmentally sound and sustainable development.
  - A goal common to all citizens of our earth.
- A goal distant from the developing world in the manner it is from the over- consuming wasteful societies of the “developed” world.

❖ **Need To Save Humanity From Extinction:**

It is incumbent upon us to save the humanity from extinction. Consequences to our activities cause destructing the environment and depleting the biosphere, in the name of development.

❖ **Need For Wise Planning of Development:**

Our survival and sustenance depend. Resources withdraw; processing and use of the product have all to be synchronized with the ecological cycles in any plan of development. Our actions should be planned ecologically for the sustenance of the environment and development.

❖ **NEED FOR PUBLICAWARENESS**

1. Growing Population: A population of over thousands of millions is growingat2.11 per cent every year. Over 17 million people are added each year. It puts considerable pressure on its natural resources and reduces the gains of development. Hence, the greatest challenge before us is to limit the population growth. Although population control does automatically lead to development, yet the development leads to a decrease in population growth rates.

2. Poverty (very Poor) : India has often been described a rich land with poor people. The poverty and environmental degradation are mixed with one another. The vast majority of our people are directly dependent on the nature resources of the country for their basic needs of food, fuel shelter and fodder. About 40% of our people are still below the poverty line.

3. Environment degradation: has adversely affected the poor who depend upon the resources of their immediate surroundings. Thus, the challenge of poverty and the challenge of environment degradation are two facets of the same challenge.

4. Agricultural Growth: The people must be made familiar with the methods to sustain and increase agricultural growth without damaging the environment. High yielding varieties have caused soil salinity and damage to physical structure of soil.

5. **Need to Increase Ground water:** It is essential of rationalizing the use of groundwater. Factors like community wastes, industrial effluents, chemical fertilizers and pesticides have polluted our surface water and affected quality of the groundwater. It is essential to restore the water quality of our rivers and other water bodies. Suitable strategies for conservation of water, provision of safe drinking water and keeping water bodies clean should be developed.

6. Development and Forests: Forests serve catchments for the rivers. With increasing demand of water, plan to harness the mighty river through large irrigation projects were made. Certainly, these would submerge forests; displace local people, damage flora and fauna. As such, the dams on the river Narmada, Bhagirathi and elsewhere have become areas of political and scientific debate. Forests in India have been shrinking for several centuries owing to pressures of agriculture and other uses. Vast areas that were once green stand today as waste lands. These areas are to be brought back under vegetative cover. The tribal communities inhabiting forests, respects the trees, birds and animals give them sustenance. We must recognize the role of these people in restoring and conserving forests. The modern knowledge and skills of the forest department should be integrated with the traditional knowledge and experience of the local Communities. The strategies for the joint management of forests should be evolved in a well- planned way.

7. Degradation of Land: At present out of the total 329 mha of land, only 266 mha possess any potential for production. Of this, 143 mha is agricultural land nearly and 85 suffer from varying degrees of soil degradation. Of the remaining 123 mha, 40 are completely unproductive. The remaining83 mha is classified as forest land, of which over half is denuded to various degrees. Nearly 406 million head of livestock have to be supported on 13 mha, or less than 4 per cent of the land classified as pasture land, most of which is over grazed. Thus, out of 226 mha, about 175 mha or 66 per cent is degraded to varying degrees. Water and wind erosion causes further degradation of almost 150 mha this degradation is to be avoided.

8. Evil Consequences of Urbanization: Nearly 27% of Indians live in urban areas. Urbanization and industrialization has given birth to a great number of environmental problems. Over 30 percent of urban Indians live in slums. Out of India's 3,245 towns and cities, only 21 have partial or full sewerage and treatment facilities. Hence, coping with rapid urbanization is a major challenge.

9. Air and water Pollution: Majority of our industrial plants are using out dated and pollution causing technologies and makeshift facilities devoid of any provision of treating their wastes. A great number of cities and industrial areas have been identified as the worst in terms of air and water pollution. Acts are enforced in the country, but their implement is not so easy. The reason is their implementation needs great resources, technical expertise, political and social will. Again the people are to be made aware of these rules. Their support is indispensable to implement these rules.

❖ INSTITUTIONS INENVIRONMENT

Managing natural resources require efficient institutions at all levels i.e. local, national, regional and global. Among the large number of institutions that deal with environmental protection and conservation, a few well-known organization include government organizations like the BSI and ZSI, and NGOs like the BNHS, WWF-1 *etc*.

Expected questions

1. Short Answer Type Question:

a) Define environment.
b) What are the components of environment?
c) What are the factors that affect the environmental condition?
d) Name biotic and a biotic component.
e) How are environment and human beings related?
f) Write down the basic needs of humans.
g) Write down the main scope of environmental science.
h) In which sphere ozone layer exists.
i) What is the difference between biotic and a biotic component withexample?
j) Define environmental science and environmental studies.

2. Long Answer type Question:

a) Environmental study is multidisciplinary in nature. Explain.
b) Discuss the public awareness towards the environmental education.
c) Enumerate the scope and importance of environmental studies.
d) Give the structure of environment.
e) Identify the basic components of environment and explain.

Natural Resources Unit-2

RENEWABLE AND NONRENEWABLE RESOURCES:

INTRODUCTION

- Natural resources can be defined as 'variety of goods and services provided by nature which are necessary for our day-to-day lives'.
- Eg: Plants, animals and microbes (living or biotic part), Air, water, soil, minerals, climate and solar energy (non- living or a biotic part).
- They are essential for the fulfillment of physiological, social, economic and cultural needs at the individual and community levels.

TYPES OF NATURALRESOURCES

They are of **two types** of resources namely **Renewable** and **Non-Renewable** Resources.

Renewable resources:

- . The resources that can be replenished through rapid natural cycles are known as renewable resource. These resources are able to increase their abundance through reproduction and utilization of simple substances.
  Ex: Plants, (crops and forests) and animals.
- Some examples of renewable resources though they do not have life cycle but can be recycled.
- Ex: Wood and wood-products, pulp products, natural rubber, fibers (e.g. Cotton, jute, animal wool, silk and synthetic fibers) and leather.
- In addition to these resources, water and soil are also classified as renewable resources.

Solar energy although having a finite life, as a special case, is considered as a renewable resource in as much as solar stocks is inexhaustible on the human scale

Nonrenewable resources: The resources that cannot be replenished through natural processes are known as non-renewable resources. These are available in limited amounts, which cannot be increased. These resources include fossil fuels (petrol, coal etc.), metals (iron, copper, gold, silver, lead, zinc etc.), minerals and salts (carbonates, phosphates, nitrates etc.).Once a non-renewable resource is consumed, it is gone forever.

- Non-renewable resources can further be divided into **two categories**, Such as-:

(A) Recyclable and (B) Non-recyclable

A) Recyclable: These are non-renewable resources, which can be collected after they are used and can be recycled. These are mainly the non-energy mineral resources, which occur in the earth's crust (Ex: Ores of aluminum, copper, mercury etc.) and deposits of fertilizer nutrients (e.g. Phosphate sock and potassium and minerals used in their natural state (asbestos, clay, mica etc.)

B) Non-recyclable: These are non-renewable resources, which cannot be recycled in any way. Ex: Fossil fuels and uranium, which provide 90 per cent of our energy requirements.

a) Natural resources and associated problems.

- The main problem associated with natural resources is unequal consumption.
- A major part of natural resources are consumed in the 'developed' world. The 'developing nations' also over use many resources because of their greater human population. However, the consumption of resources per capita (per individual) of the developed countries is up to50 times greater than in most developing countries.
- Advanced countries produce over 75% of global industrial waste and greenhouse gases.

- Energy from fossil fuels consumed in relatively much greater quantities in developed countries. Their per capita consumption of food too is much greater as well as their waste.

Forest resources:

Use and over-exploitation, deforestation, case studies, Timber extraction mining, dams and their effects on forests and tribal people.

A forest can be defined as a biotic community predominant of trees, shrubs or any other woody vegetation usually in a closed canopy. It is derived from Latin word *'foris'* means *'outside'*. India's Forest Cover is 6,76,000 sq.km (20.55% of geographic area). Scientists estimate that India should ideally have 33% of its land under forests. Today we only have 13 about 12% thus we need not only to protect our existing forests but also to increase our forest cover.

FUNCTIONS OF FOREST

• It performs very important function both to human and to nature.
• They are habitats to millions of plants, animals and wild life.
• They recycle rain water.
• They remove pollutant from air.
• They control water quality.
• They moderate temperature and weather.
• They influence soil condition and prevent soil erosion.

USES OF FOREST

There are two purposes for which a Forest can used

A. Commercial uses **B.** Ecological uses

A. The direct benefits from forests are (Commercial uses)

(a) Fuel Wood: Wood is used as a source of energy for cooking purpose and for keeping warm.
(b) Timber: Wood is used for making furniture, tool-handles, railway sleepers, matches, ploughs, bridges, boats etc.
(c) Bamboos: These are used for matting, flooring, baskets, ropes, rafts, cots etc.
(d) Food: Fruits, leaves, roots and tubers of plants and meat of forest animals form the food of forest tribes.
(e) Shelter: Mosses, ferns, insects, birds, reptiles, mammals and micro-organisms are provided shelter by forests.
(f) Paper: Wood and Bamboo pulp are used for manufacturing paper **(Newsprint, stationery, packing paper, sanitary paper)**
(g) Rayon: Bamboo and wood are used in the manufacture of rayon (yarns, artificial silk-fibres)
(h) Forest Products: Tannins, gums, drugs, spices, insecticides, waxes, honey, horns, musk, ivory, hides etc. are all provided by the flora and fauna of forests.

B. The indirect benefits from forests are (Ecological uses)

(a) Conservation of Soil: Forests prevent soil erosion by binding the soil with the network of roots of the different plants and reduce the velocity of wind and rain — which are the chief agents causing erosion.
(b) Soil-improvement: The fertility of the soil increases due to the humus which is formed by the decay of forest litter.
(c) Reduction of Atmospheric Pollution: By using up carbon dioxide and giving off oxygen during the process of photosynthesis, forests reduce pollution and purify the environment.
(d) Control of Climate: Transpiration of plants increases the atmospheric humidity which affects rainfall and cools the atmosphere.

(e) Control of Water flow: In the forests, the thick layer of humus acts like a big sponge and soaks rain water preventing run-off, thereby preventing flash-floods. Humus prevents quick evaporation of water, thereby ensuring a perennial supply of water to streams, springs and wells.

REASON FOR DEFICIENCY OF FOREST:

In India the minimum area of forest required to maintain good ecological balance is about 33% of total area. But at present it is only about 12%. So over exploitation of forest material occurs.

Over-exploitation of forests:

Since time immemorial, humans have depended heavily on forests for food, medicine, shelter, wood and fuel. With growing civilization the demands for raw material like timber, pulp, minerals, fuel wood etc. shooted up resulting in large scale logging, mining, road-building and clearing of forests.

Excessive use of fuel wood and charcoal, expansion of urban, agricultural and industrial areas and overgrazing have together led to over-exploitation of our forests leading to their rapid degradation.

Deforestation

Deforestation is the removal of a forest or stand of trees where the land is thereafter converted to a non-forest use. Examples of deforestation include conversion of forestland to farms, ranches, or urban use.

Causes of Deforestation

Main causes responsible for deforestation are as under:

- Felling of trees to meet the ever increasing demand of the cities.
- Grazing by the local cattle, goats, sheep etc. They not only destroy the vegetation but also pull out the roots of plants.
- Meeting out the growing hunger for land. It has hit the ecology of the country badly very soon India is likely to have more of wasteland than productive land.
- The increase in shifting (jhum) cultivation in North east and Orissa has also laid large forest tracts bare.
- Another major cause of deforestation has been the construction of hill roads. Road construction damages the protective vegetation cover both above and below roads.

Consequences of deforestation (or) impacts of deforestation:

1. Economic loss
2. Loss of biodiversity
3. Destructs the habitats of various species
4. Reduction in stream flow
5. Increases the rate of global warming
6. Disruption of weather patterns and global climate
7. Degradation of soil and acceleration of the rate of soil erosion.
8. Induces and accelerates mass movement / landslides.
9. Increases flood frequency, magnitude / severity.
10. Breaks the water cycle
11. Breaks the nutrient cycle

Prevention of deforestation (or) methods of conservation of forests:

1. New plants of more or less of the same variety should be planted to replace the trees cut down for timber
2. Use of wood for fuel should be discouraged.

3. Forest pests can be controlled by spraying pesticides by using aero planes
4. Forest fire must be controlled by modern techniques.
5. Over grazing by cattle must be controlled.
6. Steps should be taken by the government to discourage the migration of people into the islands from main land.
7. Education and awareness programs must be conducted.
8. Strict implementation of law of Forest conservation Act.

Case Study

Deforestation in the Himalayan region, involves clearing of natural forests and plantation of monoculture like Eucalyptus. Nutrient in the soil is poor; therefore soil losing their fertility, hence, Himalayan area facing the serious problem of desertification.

MAJOR ACTIVITIES IN FORESTS:

TIMBER EXTRACTION

Wood used for engineering purposes like building houses, making furniture is called timber. The products derived from timber have been important to many civilizations, and thus it has acquired value within these civilizations. Timber extraction results in deforestation and in the fragmentation of the last remaining forests. It harms valuable species of trees, birds and wild animals. In spite of this, it is sometimes necessary to extract timber, so as to meet the needs of a developing country. During the extraction of timber, cutting, felling and handling should be done selectively, carefully and in a planned manner, in order to save the remaining forests and biodiversity.

Effects of Timber Extraction

The major effects of timber extraction on forest and tribal people include:
1. Poor logging results in a degraded forest.
2. Floods may be intensified by cutting of trees or upstream water sheds.
3. Loss of biodiversity.
4. Climatic changes such as less rain.
5. New logging roads permit shifting cultivators to gain access to logged areas and cut the remaining trees.
6. It results in forest fragmentation which promotes loss of biodiversity because some species of plants and animals require large continuous areas of similar habitat to survive.
7. Exploitation of tribal people by the contractors.
8. Soil erosion especially on slopes occurs extensively.
9. Sedimentation of irrigation systems, floods may be intensified by cutting of trees on upstream.

Case Study-Chipko Movement

The world famous **Chipko Movement**, pioneered by **DasohliGram Swarajya Mandal** in Gopesh war brought about a general awareness about conservation of forests. The first Chipko Movement dates back to 1731, when a village woman named Amrita Bailed the Bishnoi women against the Maharajas men to prevent them from cutting trees. In this attempt to save the trees, she sacrificed her life along with the lives of her husband, three daughters and 363 people. The movement was given this name because the village women embraced or hugged the trees to stop them from being cut. In 1972, in Uttar Pradesh, the Chipko Movement was led by Bachnoi Devi of Advani who protected the hill forests from the contractors axe men.

DAMS:-

Today there are more than 45,000 large dams around the world, which play an important role in communities and economies that harness these water resources for their economic development. Current estimates suggest some 30-40% of irrigated land worldwide relies on dams. Hydropower, another important the use of stored water, currently supplies 19% of the world's total electric power supply and is used in over 150 countries. The world's two most populous countries – China and India –have built around 57% of the world's large dams.

Dams problems

Dams are the massive artificial structures built across the rivers to store water for much beneficial purpose.
Dams are considered a "Temples of modern India". Dams destruct vast area of forest area. India has more than 1600 large dams.

Effects of dams on forest:

• Thousands of hectares of forest will be cleared.
• Killing of wild animals and destruction of aquatic life.
• Spreading of water borne diseases.
• Water logging increases the salinity of the soil.

Ex: Narmada Sagar project it has submerged 3.5 lakhs hectares of forest.

Effects of dam on tribal people

1. Construction of big dams leads to the displacement of tribal people.
2. Displacement and cultural change affects the tribal people both mentally and physically.
3. They do not accommodate the modern food habits and lifestyle.
4. Tribal people are ill-treated by the modern society.
5. Many of the displaced people were not recognized and resettled or compensated.
6. Body condition of tribal people will not suit with new areas and hence they will be affected by many diseases.

Case study- Sardar Sarovar Project:

The World Bank's withdrawal from the Sardar Sarovar Project in India in 1993 was a result of the demands of local people threatened with the loss of their livelihoods and homes in the submergence area. This dam in Gujarat on the Narmada has displaced thousands of tribal folk, whose lives and livelihoods were linked to the river, the forests and their agricultural lands. While they and the fishermen at the estuary have lost their homeland, rich farmers downstream will get water for agriculture. The question is why should the local tribal's be made homeless, displaced and relocated to benefit other people? Why should the less fortunate be made to bear the costs of development for better off farmers? It is a question of social and economic equity as well as the enormous environmental losses, including loss of the biologicaldiversity of the inundated forests in the Narmada valley.

MINING:-

"The process of extracting mineral resources and fossil fuels like coal from the earth is called as mining."

Types of mining

1. Surface mining: Mining of minerals from shallow deposits
2. Underground mining: Mining of minerals from deep deposits

Steps involved in mining

1. Exploration
2. Development
3. Exploitation
4. Ore processing
5. Extraction and purification of minerals

The extent of damage by underground mining is more than that of surface mining, which needs enormous amount of land area for its operation and management.

Effects of mining

1. Pollute soil, water and air.
2. Destruction of natural habitat.
3. Continuous removal of minerals leads to the formation of trench where water is logged which contaminates the groundwater.
4. Vibrations cause earthquakes.
5. Produces noise pollution
6. Reduces shape and size of the forest.
7. Increased risk of landslides.
8. Spoils the aesthetic beauty.

Water resources: Use and over-utilization of surface and ground water, floods, drought, conflicts over water, dam's benefits and problems.

Water is the most abundant, inexhaustible renewable resource. It covers 70% of the globe in the form of oceans, rivers, lakes, etc. Of this 70%, only 3% is available as freshwater. From this 3%, roughly 2% is frozen in polar icecaps and only a fraction of the remaining 1% is used as drinking water (potable). 90% of the water is utilized for agricultural purposes in India.

USE OF SURFACE AND GROUND WATER

Consumptive use: In such uses, water is completely utilized and cannot be reused. Ex: Domestic, industrial and irrigation

Non-consumptive use: In such uses, water is not completely utilized and is reused Ex: Hydropower plant

Other uses:

- Water is used for domestic purposes like drinking, bathing, cooking, washing. etc.
- Water is used in commercial establishments like hotels, theaters, educational institutions, offices, etc.
- Almost 60-70% of fresh water is used for irrigation
- 20-30% of water is used for industrial operations by refineries, iron & steel industries, paper & pulp industries, etc.
- Water plays a key role in sculpting the earth's surface, moderating climate and diluting pollutants.

Over-utilization of surface & ground water

- The rapid increase in population and industrial growth led to severe demand on water resources. After using all available surface water resources to the maximum, human beings began using groundwater to meet their needs.
- The increased extraction of groundwater far in excess of the natural recharge led to decreased groundwater level. The erratic and inadequate rainfall caused reduction in storage of water in reservoirs. This also led to decrease of

groundwater.

- Building construction activities seal permeable soil zone and reduce the area for percolation of rainwater thereby increasing surface runoff.
- If groundwater withdrawal rate is higher than recharge rate, sediments in aquifers get compacted resulting in sinking of overlaying land surface. This is called land subsidence which leads to structural damage in buildings, fracture in pipes and reverses the flow of canals leading to tidal flooding.
- Over-utilization of groundwater in arid and semi-arid regions for agriculture disturbs equilibrium of reservoir in the region causing problems like lowering of water table and decreased pressure in aquifers coupled with changes in speed and direction of water flow.
- Over utilization of groundwater in coastal areas leads to rapid intrusion of salt water from the sea thereby rendering it unusable for drinking and agriculture.
- Over-utilization of groundwater lads to decrease in water level thereby causing earthquake, landslides and famine.
- Over-utilization of groundwater leads to drying-up of dug wells as well as bore wells.
- Due to excess use of groundwater near agricultural fields, agricultural water that contains nitrogen as a fertilizer percolates rapidly and pollutes the groundwater thereby rendering the water unfit for potable use by infants. (Nitrate concentration exceeding 45 mg/L).

FLOOD:-

It is an over flow of water. It happens when the magnitude of flow of water exceeds the carrying capacity of the channel within its bank.

CAUSES OFFLOOD

- Heavy rainfall, melting of snow and sudden release of water from dams. (Flashfloods)
- Reduction in the carrying capacity of the channel.
- Deforestation, mining and over grazing increase the runoff from rains and the level of flood raises.

EFFECT OFFLOOD

- Water spreads in the surrounding area and submerges them.
- Cultivated land gets affected.
- Extinction of civilization.

FLOODMANAGEMENT

- Floods can be controlled by dams.
- Channel management control flood.
- Flood hazards reduced by forecasting or flood warning.
- Flood may also be reduced by reduction of run off by increasing infiltration through appropriate a forestation in the catchment area.

DROUGHT:-

Drought is nothing but scarcity of water, which occurs due to

- Inadequate rainfall
- Late arrival of rainfall
- Excessive withdrawal of groundwater.

- Lack of water for the needs of agriculture, livestock, industry or human population may be termed as a drought. Drought causes serious damages to plants, animals and human life.

CAUSES OF DROUGHT

- When annual rain fall is below normal and less than evaporation, drought is created.
- High population.
- Intensive cropping pattern
- Ex: Maharashtra - There has been no recovery from drought for the last 30 years due to over exploitation of water by sugarcane crop.

EFFECTS OF DROUGHT

- Drought causes hunger, malnutrition and scarcity of drinking water an also changes the quality of water.
- Drought causes widespread crop failure leading to acute shortage of food and adversely affects human and livestock population.
- Worst situation of drought causes desertification.
- Raw materials of agro based industries are critically affected during drought time, hence industrial and commercial growth decreases.
- Drought increases the degradation of natural resources.
- Drought causes large migration of people and urbanization.

DROUGHT MANAGEMENT

- Indigenous knowledge is essential.
- Rain water harvesting system.
- Construction of reservoirs to improve ground water level.
- Modern irrigation technology (drip irrigation) very useful to conserve water.
- Forestation activities also improve the potential of water in the drought area.
- Crop mixing and dry forming are the suitable methods which minimize the risk of crop failures in dry area.

DAMS:-

Dams made significant contributions to human development and the benefits derived from them have been considerable. Large dams are designed to control floods and to help the drought prone areas, with supply of water. But large dams have proved to cause severe environmental damage. Hence an attempt has been made to construct small dams. Multiple small dams have less impact on the environment.

Benefits: Dams ensure a year round supply of water for domestic use and provide extra Water for agriculture, industries and hydropower generation.

Problems: They alter river flows, change nature's flood control mechanisms such as wetlands and flood plains, and destroy the lives of local people and the habitats of wild plant and animal species, particularly are the case with mega dams. Some of the problems are mentioned below.

- Dam construction and submersion leads to significant loss of farmland and forest and land submergence
- Siltation of reservoirs, water logging and salination in surrounding lands reduces agricultural productivity
- Serious impacts on ecosystems - significant and irreversible loss of species and ecosystems, deforestation and loss of biodiversity, affects aquaculture

- Socio economic problems for example, displacement, rehabilitation and resettlement of tribal people.
- Fragmentation and physical transformation of rivers
- Displacement of people - People living in the catchment area, lose property and livelihood
- Impacts on lives, livelihoods, cultures and spiritual existence of indigenous and tribal people
- Dislodging animal populations
- Disruption of fish movement and navigational activities
- Emission of greenhouse gases due to rotting of vegetation
- Natural disasters – reservoirs induced seismicity, flash floods etc. and biological hazards due to large-scale impounding of water – increase exposure to vector bone diseases, such as malaria, schistosomiasis, and filariasis.

SUSTAINABLE WATERMANAGEMENT

- Building several small reservoirs instead of few megaprojects
- Developing small catchment dams and protecting wetlands
- Water conservation measures in agriculture, such as using drip irrigation, control of growing water intensive cash crops; control of water logging.
- Effective rainwater harvesting in urban environments
- Treating and recycling municipal waste water for agricultural use.
- Soil management, micro-catchment development and a forestation permits recharging of underground aquifer, thus reducing the need for large dams
- Preventing leakages form dams and canals and loss in municipal pipes
- Pricing water at its real value makes people use it more responsibility and efficiently and reduces wastage
- In deforested areas where land has been degraded, appropriate soil management practices, making bunds along the hill-slopes and making nalla plugs can help retain moisture and make it possible to vegetate degraded areas
- Use waste water for activities that does not need fresh water –Recycling
- Adopt mini water harvesting models for domestic usage.
- Protect existing tanks
- Develop systematic water management and adopt strict water auditing
- “Save water Campaigns” for public awareness on water scarcity
- Through rainwater harvesting, community based participatory initiatives and holistic watershed management.
- Responsible water usage can only be achieved by empowering local communities and creating local accountability.
- The government should develop policies that protect water resources, promote sustainable watershed management and invest in technologies that will increase efficiency in irrigation, industrial usage and improve water harvesting techniques.

WATERCONFLICTS

1. Conflict through use: Unequal distribution of water led to interstate and international disputes.

National conflicts:

- Sharing of Cauvery water between Karnataka and Tamil Nadu.
- Sharing of Krishna water between Karnataka and Andhra Pradesh
- Siruvani– Tamil Nadu and, Kerala

International conflicts:

Indus – India and Pakistan &Colorado River – Mexico and USA Mineral

Mineral Resources: Use and exploitation, environmental effects of extracting and using mineral resources.

Resources: Use and exploitation, environmental effects of extracting and using mineral resources.

USE

Mineral Resources in India such as coal, mineral oil, iron-ore, gypsum, limestone, dolomite, mica, bauxite etc. are used for the following purposes.

Coal: Coal has 1 lakhs 20 thousand bi-products like coke, coke gas, Varnish, Colour and Paint, Cement, Fertilizer, Explosive, Nylon, Terylin, Asphalt, Naphtha, Thermocol, Thermal Electricity, Saccarine, Benzal, Polish, Ink etc.

Mineral Oil: It has more than 80 thousand bi-products: Petrol, kerosene, fuel oil, lubricating oil, diesel, nylon, terylin, color and paints, pesticides, explosive, chemicals, plastics, fertilizer.

Iron-ore: Mainly used in iron and steel, Ship-building, Automobile, Railway, Aircraft, Heavy Machine tools, Gypsum: It is mainly used in chemical industries especially in fertilizer and other chemical products.

Limestone and dolomite: Mainly used in Iron & steel, Paper, House-building, Color and Paints, Chemicals, Fertilizer etc.

Mica: Mainly used in electrical goods, decoration, colour and paints, aeronautical engineering, motor transport, medicinal preparation, ornament etc.

Bauxite: Used in aeronautical engineering, chemicals, color and paints, electrical goods, railway coach, Bus and Motor Coach, Electrical wires, Utensils, Furniture, Doors and windows, Coins and Image-making Industry.

Copper: Mainly used in Electrical goods, making Engineering parts, Colors and Paints, Chemicals, Image making, Coins, Telephone etc.

Manganese: Mainly used in iron and steels, and chemical industry, enamel power, glass and electrical equipment's.

Exploitation of Mineral Resources

Exploitation of mineral refers to the use of mineral resources for economic growth. Exploitation of mineral resources at a mindless speed to meet the growing needs of modern civilization has resulted in many environmental problems. Today, about 80% of the world's energy consumption is sustained by the extraction of fossil fuels, which consists of oil, coal, and gas.

ENVIRONMENTAL EFFECTS OF EXTRACTING AND USING MINERAL RESOURCES:-

(a) Mining is hazardous occupation:

1. This occupation involves several health risk dust produced during mining operation are injurious to health and cause lung diseases.
2. Extraction of some toxic or radioactive minerals leads to life threatening hazards.
3. Dynamite explosion during mining is very risky as fumes produced are extremely poisonous.
4. Underground mining is more hazardous than surface mining as there are more chances if accidents like roof falls, flooding and inadequate ventilation etc.

(b) Rapid depletion of high grade minerals:

Increasing demand for high grade minerals has compelled miners to carry out more extraction of minerals, which require more energy sources and produce large amount of waste materials.

(c) Wastage of upper soil layer and vegetation:

Surface mining results in the complete destruction of upper soil layer and vegetation. After extraction, the wastes are dumped in an area which destroys the total surface and vegetation.

(d) Environmental problems:

Over exploitation of mineral resources resulted in many environmental problems like:

1. Conversion of productive land into mining and industrial areas.
2. Mining and extraction process are one of the sources of air, water and land pollution.
3. Mining involves huge consumption of energy resources like coal, petroleum, natural gas etc. which are in-turn nonrenewable sources of energy.
4. Surface mining directly degrades the fertile soil surface thus effect ecology and climate if that particular area.

CASE STUDIES-MINING AND QUARRYING INUDAIPUR

- 200 open cast mining and quarrying in Udaipur. But 100 mines are illegal. 150 tons of explosives are used per month. It pollutes air, soil and water. It affects irrigation and wild life.

{**Food Resources:** World food problems, changes caused by agriculture and over grazing, effects of modern agriculture, fertilizers- pesticides problems, water logging, salinity }

- Food is essential for growth and development of living organisms. These essential materials are called nutrients and these nutrients are available from variety of animals and plants. There are thousands of edible plants and animals over the world, out of which only about three dozen types constitute major food of humans**.**
- Food is an essential requirement for survival of life. Main components are carbohydrates, fats, proteins, minerals and vitamins.

TYPES OF FOODSUPPLY

- **Crop plants:** Grains mostly **c**onstitute about 76% of the world's food. Ex: Rice, Wheat and Maize
- **Range lands: P**roduces 17% of world's food from trees and grazing animals.
  Ex: Fruits, milk and meat
- **Ocean:** Fisheries – 7% of world's food

WORLD FOOD PROBLEM :-

1. In the earth's surface, 79% is water out of total area. 21% land (forest, desert, mountain and barren land). Less % cultivated land, at the same time population explosion is high therefore world food problem arises.
2. Environmental degradation like soil erosion, water logging, water pollution, salinity affects agricultural land.
3. Urbanization affects agricultural land. Hence production of rice, wheat, corn and other vegetable is difficult.
4. Poor agricultural practices: Poor environmental agricultural practices such as slash and burn, shifting cultivation, or 'rab' (wood ash) cultivation degrade forests.
5. Degradation of agricultural lands: Globally 5 to 7 million hectares of farmland is degraded each year. Loss of nutrients and overuse of agricultural chemicals are major factors in land degradation.
6. Water scarcity is an important aspect of poor agricultural outputs. Salinization and water logging has affected a large amount of agricultural land worldwide.
7. Our fertile soils are being exploited faster than they can recuperate.
8. Forests, grasslands and wetlands have been converted to agricultural use, which has led to serious ecological questions.
9. Use of genetically modified seed variety, without minding the conducive environment for such experimentation,

will seriously affect the land ecosystem.

10. Our fish resources, both marine and inland, show evidence of exhaustion.
11. There are great disparities in the availability of nutritious food. Some communities such as tribal people still face serious food problems leading to malnutrition especially among women and children.
12. Loss of Genetic Diversity: Modern agricultural practices have resulted in a serious loss of genetic variability of crops. India's distinctive traditional varieties of rice alone are said to have numbered between 30 and 50 thousand. Most of these have been lost to the farmer during the last few decades as multinational seed companies push a few commercial types.

TYPES OFNUTRITION

• Nutritious nutrition: To maintain good health and disease resistance, we need large amount of carbohydrate, proteins, fats and smaller amount of micronutrients such as vitamins and minerals such as Fe, Ca and iodine. Food and agricultural organization (FAO) of United Nations estimated that on an average, the minimum calorie intake on a global state is 2500 calories/day.

• Under nutrition: People who cannot buy enough food to meet their basic energy needs suffer from under nutrition. They receive less than 90% of this minimum dietary calorie. Effect of under nutrition: Suffer from mental retardation and infectious diseases.

• Mal nutrition: Besides minimum calorie intake we also need proteins, minerals, vitamins, iron and iodine. Deficiency leads to malnutrition resulting in several diseases.

- India 3rd largest producer of crops, nearly 300 million Indians are still under nourished.

World food summit 1996: The world food summit, 1996 has set the goal to reduce the number of under nourished and mal nourished people to just half by 2015.

Changes Caused by Agriculture and Overgrazing

From centuries, agriculture is providing inputs to large number of industries involved in production, processing and distribution of food. Accordingly, agriculture has significant effect on environment. The effects of agriculture on environment can be classified as local, regional, and global level. The agriculture also makes impact on the usage of land generally as follows:

1. Deforestation
2. Soil Erosion
3. Depletion of nutrients
4. Impact related to high yielding varieties (HYV)
5. Fertilizers related problems include micronutrient imbalance, nitrite pollution and eutrophication.
6. Pesticide related problems include creating resistance in pests and producing new pests, death of non-target organisms, biological magnification.
7. Some other problems include water logging, salinity problems and such others.
8. The carrying capacity of land for cattle depends upon micro climate and soil fertility. If carrying capacity is exceeded than land is overgrazed. Because of overgrazing the agricultural land gets affected as follows.
9. Reduction in growth and diversity of plant species.
10. Reduce plant cover leads to increased soil erosion.
11. Cattle trampling leads to land degradation.

TYPES OF AGRICULTURE

(1) Traditional agriculture **(2) Modern (or) industrialized agriculture**

1. Traditional agriculture

Small plot, simple tools, surface water, organic fertilizer and a mixture of crops constitute traditional agriculture. They produce enough food to feed their family and to sell it for their income.

2. Modern agriculture

Hybrid seeds of single crop variety, high tech equipment, lot of fertilizers, pesticides and water to produce large amount of single crops.

Effects of Modern Agriculture

For sustainable production modern techniques are used to enhance productivity of different cropping systems under different agro-eco-zones. Adoption of modern agricultural practices has both positive and negative effects on environment. Effects of modern agriculture are briefly discussed under different heads as under:

- **Soil erosion**

Raindrops bombarding bare soil result in the oldest and still most serious problem of agriculture. The long history of soil erosion and its impact on civilization is one of devastation.

- **Irrigation**

Irrigation ensures sufficient water when needed and also allows farmers to expand their acreage of suitable cropland. In fact, we rely heavily on crops from irrigated lands, with fully one-third of the world's harvest coming from that 17% of cropland that is under irrigation. Unfortunately, current irrigation practices severely damage the cropland and the aquatic systems from which the water is withdrawn.

- **Agriculture and the loss of genetic diversity**

As modern agriculture converts an ever-increasing portion of the earth's land surface to monoculture, the genetic and ecological diversity of the planet erodes.

- **Fertilizer-pesticide problems**

On one hand application of artificial chemical fertilizers increases the productivity at faster rate as compare to organic fertilizers, on the other hand application of fertilizers can be a serious problem of pollution and can create number of problems.

- Excessive level of nitrates in ground water has created problems in developed countries.
- Accumulated phosphorous as a consequence of use of phosphoric fertilizer are posing serious threat as residues in domestic water supply and for ecology of river and other water bodies.
- Increased level of phosphates in different water results in eutrophication.
- Effect of chemical fertilizer is long term, therefore leads to net loss of soil organic matter.
- To control insects, pests, diseases and weeds which are responsible for reduction in productivity different chemicals are used as insecticides, pesticides and herbicides. Applications of these synthetic chemicals have great economic values and at the same time cause number of serious problems such as:
- Affects human health which includes acute poisoning and illness caused by higher doses and accidental exposes
- As long term effect, cause cancer, birth defects, Parkinson's disease and other regenerative diseases.
- Long term application of pesticides can affect soil fertility.
- Danger of killing beneficial predators.
- Pesticides resistance and pest resurgence

Water Logging

High water table or surface flooding can cause water logging problems .Water logging may lead to poor crop productivity due to anaerobic condition created in the soil. In India, deltas of Ganga, Andaman and Nicobar Islands and some areas of Kerala are prone to frequent water logging.

CAUSES OF WATER LOGGING

- Excessive water supply to the croplands
- Heavy rain
- Poor drainage

MEASURES TO PREVENT WATER LOGGING

- Avoid and prevent excessive irrigation
- Sub-surface drainage technology
- Bio-drainage by trees like Eucalyptus

Salinity

- Due to adoption of intensive agriculture practices and increased concentration of soluble salts leads to salinity. Due to poor drainage, dissolved salts accumulate on soil surface and affects soil fertility. Excess concentration of these salts may form a crust on the surface which may injurious to the plants. The water absorption process is affected and uptake of nutrient is disturbed. According to an estimate, in India, 7 million hectare of land is saline and area is showing in increasing trends due to adoption of intensive agriculture practices.

Case Studies

- Food Centre at Center for Science and Environment (CSE) India reported Pepsi and Coca-Cola companies sold soft drinks with pesticide content 30-40 times higher than EU guidelines permit. At the reported concentrations the pesticides damage the nervous system
- Recent reports from cotton growing belt of Punjab which covers Abohar, Fazalka and part of Bathinda indicates that over use of pesticides for control of insect pest in cotton to enhance productivity has not only affected soil health, but also caused cancer in human being.
- In Delhi, accumulation of pesticides and DDT in the body of mothers caused premature deliveries or low birth weight infants.
- Canal irrigation in Haryana resulted in rising water table followed by water logging and salinity causing low crop productivity thereby huge economic losses.

  [Energy Resources: Growing energy need, renewable and non-renewable energy sources, use of alternate energy sources, case studies]

ENERGY DISTRIBUTION IN THE WORLD

- Developed countries like USA and Canada constitute only 5% of the world's population but consume 25% of the world's available energy.
- Energy consumed by a person in a developed country for a single day is equal to energy consumed by a single person in a poor country for one year.
- Developed country GNP increases and energy consumption increases. In the poor country GNP and energy consumption are less.

TYPES OF ENERGY RESOURCES:

1. Renewable energy resource (or) Non-conventional energy resources.

2. Nonrenewable energy resources (or) Conventional energy resources.

1. RENEWABLE ENERGY SOURCES:

Energy which can be regenerated or these resources can be generated continuously and are inexhaustible.

Ex: Wood, Solar energy, Wind energy, Hydro power, tidal energy, Geo-thermal energy, etc.

Merits of renewable energy resources

- Unlimited supply
- Provides energy security.
- Fits into sustainable development concept.
- Reliable and the devices are modular in size.
- Decentralized energy production.

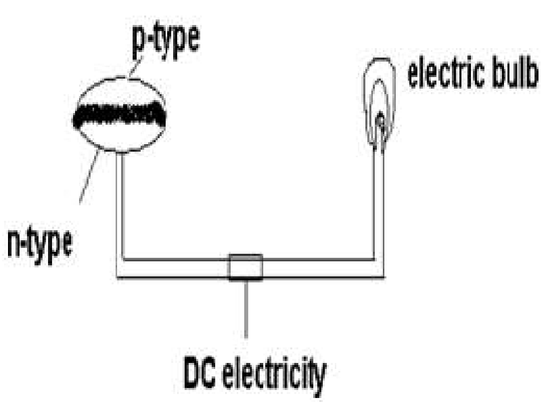
Types of renewable energy resources

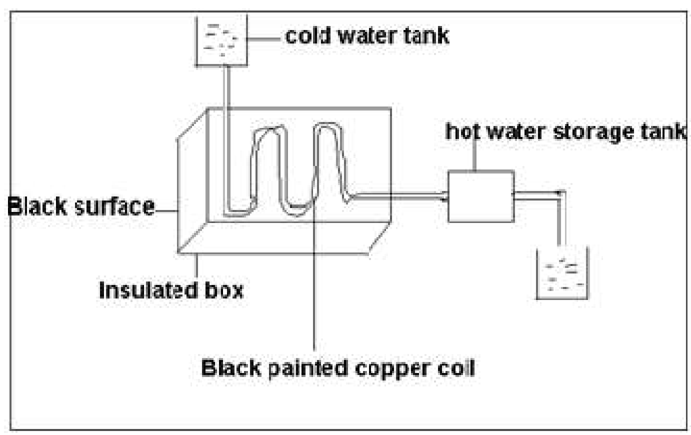
Solar energy: Nuclear fusion reaction of sun produces enormous amount of energy. Several techniques are available for collecting, storing and using solar energy.

Solar cell (or) Photovoltaic cell (or) PV cell:

- Solar cell consists of p- type semiconductor (Si doped with B) and n-type semiconductor (Si doped with P). P-type forms top layer and n-type forms bottom layer.
- Solar rays fall on the top layer, the electrons from valence band promoted to the conduction band which crosses the p-n junction into n-type semiconductor. Potential difference between the two layers is created which causes flow of electrons.

  Uses: It is used in calculators, electronic watches, street light, water pump sets.





A. Solar battery: Large number of solar cells connected in series is called solar battery. It is used in remote areas where continuous power supply is a problem.

B. Solar water heater: It consists of insulated box painted with black paint with glass lid. Inside the box black painted copper coil is present. Cold water is allowed to flow, it is heated up and flows out into a storage tank from which water is supplied through pipes.

Wind energy: Moving air is called wind. The energy recovered from the force of the wind is called wind energy its speed is high.

a. Wind mills: When a blowing wind strikes the blade of the windmill, it rotates continuously. And rotational motion of the blade drives number of machines like water pump, flour mills and electric generators.

b. Wind farms: When a large number of mills are installed and joined together in a definite pattern it forms wind farm. It produces large amount of electricity.

Condition: Minimum speed for wind generator is 15 Km/hr.

Advantages:

- It does not cause air pollution
- Very cheap

Ocean energy:

Tidal energy (or) Tidal power: Ocean tides are due to gravitational force of sun and moon which produce enormous amount of energy. High tides – rise of water in the ocean. Low tides – fall of water in the ocean. Tidal energy can be used by constructing a tidal barrage. During high tides sea water enters into the reservoirs and rotates the turbine, produce electricity. During low tides water from reservoir enters into the sea rotate the turbine produce electricity.

Ocean thermal energy:

- Temperature difference between surface water and deeper level water in ocean generates electricity. The energy available due to the difference in temperature of water is called ocean thermal energy.

Condition: Temperature difference should be 200C.

Process: Ammonia is converted into vapors on the surface of warm water; it increases the vapor pressures which rotate the turbine and generates electricity. Deeper level cold water is pumped to cool and condense the vapor in to liquid.

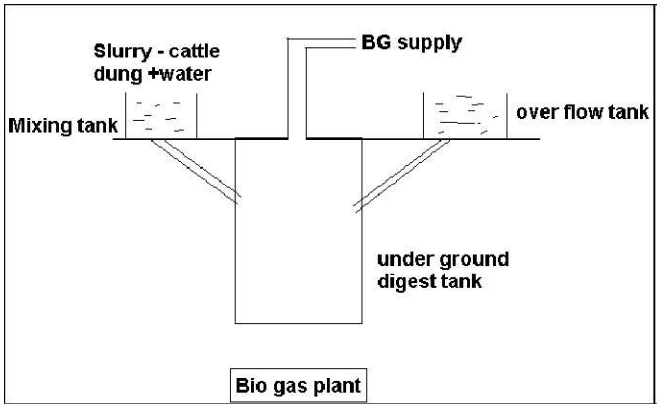
Geo thermal energy:

Temperature of the earth increase sat a of 20 –750C per/km when we move down the earth. The energy utilized from the high temperature present inside the earth is called geo thermal energy.

- **Natural geysers:** Hot water or steam comes out of the ground through cracks naturally is called natural geysers.
- **Artificial Geysers:** Artificially a drill hole up to the hot region and by sending a pipe into it. The hot water or steam is used to rotate the turbine and generate electricity.

Bio Mass energy:

- **Bio mass:** Organic matter produced by plants or animals used as source of energy
- **Bio gas:** Mixture of methane, carbon dioxide and hydrogen sulphide. Methane is the major constituent. It is obtained by anaerobic fermentation of animal dung (or) plant wastes in the presence of water.
- **Bio fuels:** Fuels obtained by the fermentation of biomass. Ex: Ethanol, methanol
- **Ethanol:** Produced from sugar cane. Calorific value is less.
- **Methanol:** Obtained from ethanol Calorific value too less.
- **Gasohol:** Mixture of ethanol and gasoline India trial is going on to use gasohol in cars and buses.
- **Hydrogen fuel:** Hydrogen produced by pyrolysis, photolysis and electrolysis of water. It has high calorific value. Nonpolluting one because the combustion product is water.



Disadvantages:

• Hydrogen is highly inflammable and explosive.
• Safe handling is required.
• Difficult to store and transport.

2. NON RENEWABLE ENERGYSOURCES:

Energy which cannot be regenerated is called as non-renewable.

Coal: It is a solid fossil fuel.

Disadvantages:

• When coal is burnt large amount of CO_2 is released which causes global warming.
• S, N produces toxic gases during burning.

Petroleum: Crude oil is a liquid consists of more than hundreds of hydrocarbons and small amount of impurities. The petroleum can be refined by fractional distillation. In the world level 25% of oil reserves are in Saudi Arabia. At present rate of usage, the world crude oil reserves are expected to get exhausted in just 40years.

Liquefied petroleum gas (LPG): Petroleum gases obtained during FD and cracking can be easily converted into liquid under high pressure as LPG. It is colorless and odorless gas, but during cylindering mercaptans are added to detect leakage.

Natural gas: These are found above oil in oil wells. It is a mixture of methane and other hydrocarbons. Calorific value is high. There are two types. Dry gas and wet gas.

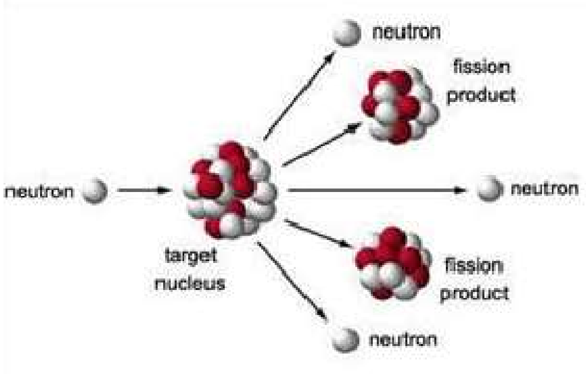
Nuclear energy: Dr.H.Bhabhais a father of nuclear power development in India. 10 nuclear reactors are present in India. It produces 2% of India's electricity. Nuclear energy can be produced by two types of reactions. Nuclear fission and nuclear fusion.

Nuclear fission: is a nuclear change in which heavier nucleus split into lighter nuclei on bombardment of fast moving neutrons. Large amount of energy is released through chain reaction.

Ex: Uranium with fast moving neutron gives barium and krypton in addition to three neutrons; in the second stage it gives nine neutrons and so on. This process of propagation of the reaction by multiplication is called chain reaction.

Nuclear fusion: It is a nuclear change in which lighter nucleus is combined together at extremely high temperature (1 billion 0C) to form heavier nucleus and a large amount of energy is released.

Ex: Isotopes of hydrogen combine to form helium molecule.



Tidal Power

The surface of earth is 71.11% covered by water bodies especially oceans. The tides in water rise and fall due to the gravity of sun and moon. Since we know about how the position of moon changes we can predict the rise and fall of tides. This rise and fall of tides can be utilized by setting up small dams and passing water through the turbines to generate power.

Advantages of tidal energy

- The source of power generation is free and renewable.
- The power generated is clean and does not cause any pollution.

CASE STUDY

- **Wind energy in India:** India generating 1200 MW electricity using the wind energy. Largest wind farm situated near Kanya kumari in Tamilnadu. It produces 380 MW electricity.
- **Hydrogen fuel car:** General motor company of china discovered a experimental car (fuelH2) can produce no emission only water droplets and vapors come out of the exhaust pipe. This car will be commercially available by2010.
- **Land Resources:** Land as a resource, land degradation, man induces landslides, soil erosion, and desertification.
- Land is a very valuable resource. It provides food, fiber, wood, medicine and other biological materials needed for food. Soil is a mixture of inorganic materials and (rocks and minerals) and organic materials (dead materials and plants).
- Top soil is classified as a renewable resource as it is continuously regenerated by natural processes at a very slow rate. However, if the rate of erosion is faster than the rate of renewal, the soil becomes a non-renewable resource.

Uses of land resources

- Land provides food, wood, minerals, etc.
- Land nurtures plants and animals that provide us food and shelter
- Land may be used as watershed or reservoir.
- Land acts as a dustbin for the wastes generated by modern society.
- Land is used for constructing buildings and industries.
- Land degradation- Landslides, Soil erosion and Desertification

LAND DEGRADATION

Land degradation is the process of deterioration of soil or loss of fertility of soil.

EFFECTS OF LAND DEGRADATION

- Soil texture and structure are deteriorated
- Loss of soil fertility due to loss of valuable nutrients
- Increase in water logging, salinity, alkalinity and acidity problems
- Loss at a social, economic and biodiversity level

CAUSES OF LAND DEGRADATION:-

- **Population:** With the increase in population, more land is needed for producing food, fibre and fuel wood leading to increasing pressure on the limited land resources. Therefore the land gets degraded due to over exploitation
- **Urbanization:** Increased urbanization due to population growth reduces the agricultural land. To compensate for loss of agricultural land, new lands comprising of natural ecosystems such as forests are cleared. Therefore, urbanization leads to deforestation which in-turn affects millions of plant and animal species.
- **Fertilizers and Pesticides:** Increased application of fertilizers and pesticides are needed to increase farm output in new lands thereby leading to pollution of land, water and soil degradation.
- **Damage to top soil:** Increase in food production generally leads to damage of top soil through nutrient depletion.
- Water-logging, soil erosion, salination and contamination of the soil with industrial waste cause land degradation.

LANDSLIDES

- Landslides are the downward movement of a slope composed of earth materials such as rock, soil or artificial fills. Landslides are also called rock-slide, debris-slide, slump, earth-flow or soil-creep.
- During construction of roads and mining activities huge portions of mountainous fragile areas are cut down and thrown into adjacent areas and streams. These land masses weaken the already fragile mountain slopes leading to man-induced landslides.

EFFECTS OF LANDSLIDES:

- Landslides increase the turbidity of nearby streams, thereby reducing their productivity
- Destruction of communicative links
- Loss of habitat and biodiversity
- Loss of infrastructure and economic loss

CAUSES OF LANDSLIDES

1. Removal of vegetation - Deforestation in slopes creates soil erosion leading to landslides.
2. Underground mining activities cause subsidence of the ground
3. Movement of heavy vehicles in areas with unstable slopes causes landslides.
4. Addition of weight by construction on slopes causes landslides.
5. Over exploitation of groundwater also leads to landslides.

SOIL EROSION

Soil erosion is the process of removal of superficial layer of soil. Soil erosion removes soil components and litter.

Harmful effects of soil erosion

1. Soil fertility is lost
2. Loss of soil ability to hold water and sediment
3. Sediment runoff can pollute water courses and kill aquatic life

Types of soil erosion

1. Normal erosion: This is caused by the gradual removal of topsoil by natural processes. The rate of erosion is slow.

2. Accelerated erosion: This is caused by man-made activities. In this case, the rate of erosion is much faster than the rate of formation of soil.

CAUSES OF SOIL EROSION

1. **Water:** Water affects soil erosion in the form of rain, run-off, rapid flow or wave action
2. **Wind:** Wind is an important climate agent that carries away the fine particles of soil thereby contributing to soil erosion.
3. **Biotic agents:** Overgrazing, mining and deforestation are the major biotic agents causing soil erosion. These processes disturb the top soil thereby exposing the soil to various physical forces inducing erosion
4. Landslides cause soil erosion
5. Construction of dams, buildings and roads removes the protective vegetal cover leading to soil erosion

SOIL CONSERVATION PRACTICES

1. **Conservational till farming or no-till farming:** Traditionally, land is ploughed to make a planting surface. This disturbs the soil and makes it susceptible to erosion. The no-till farming method makes minimum disturbance to the top soil by making slits in the unplugged soil. Seeds, fertilizers and water are injected in these slits.
2. **Contour farming:** In this method, crops are planted in rows along contours of gently sloped land. Each row acts as a small dam to hold soil thereby slowing water runoff.
3. **Terracing:** In this method, steep slopes are converted into a series of broad terraces that run across the contour. This retains water for crops and reduces soil erosion by controlling runoff.

4. Alley cropping or Agro forestry: This method involves planting crops in strips or alleys between rows of trees or shrubs that provide fruits and fuel wood. Hence, when the crop is harvested, the soil will not be eroded as the trees and shrubs remain on ground holding the soil particles.

5. Wind breaks or shelter belts: In this technique, trees are planted in long rows along the boundary of cultivated land which block the wind and reduce soil erosion. Wind breaks help in retaining soil moisture, supply wood for fuel and provide habitat for birds.

DESERTIFICATION

1. Desertification is a progressive destruction or degradation of arid or semi-arid lands to desert.
2. Desertification leads to conversion of range-lands or irrigated croplands to desert like conditions in which agricultural productivity falls.
3. Desertification is classified by de-vegetation, depletion of groundwater, salination and soil erosion.

EFFECTS OF DESERTIFICATION

Almost 80% of the productive land in the arid and semi-arid regions is converted into desert.
Approximately 600 million people are threatened by desertification.

CAUSES OF DESERTIFICATION

1. **Desertification:** Lack of vegetation prevents the rainfall from soaking into the ground resulting in poor recharge of groundwater. Eventually this results in soil erosion and loss of fertility.
2. **Over-grazing:** Increase in cattle population coupled with repeated grazing at the same location results in depletion of vegetation in the area. Eventually, the land becomes loose and prone to soil erosion and formation of a desert.
3. **Water management:** Over-utilization of groundwater, particularly in the coastal regions, results in saline water intrusion into aquifers thereby making water unfit for irrigation.
4. **Mining and quarrying:** These activities are responsible for loss of vegetative cover and denudation of extensive land area leading to desertification.
5. **Climate change:** Climate change manifests in the form of failure of monsoons, irregular monsoons and frequent droughts thereby leading to desertification
6. **Pollution:** Excessive use of fertilizers and pesticides to increase yield and disposal of toxic wastes into land leads to desertification.

b) Role of individual in conservation of natural resources.

Role of an individual in conservation of natural resources Conservation of energy:

1. Switch off light, fan and other appliances when not in use.
2. Use solar heater for cooking.
3. Dry the cloth in the sun light instead of driers.
4. Use always pressure cookers
5. Grow trees near the house to get cool breeze instead of using AC and air cooler.
6. Ride bicycle or just walk instead of using scooter for a short distance.

Conservation of water:

1. Use minimum water for all domestic purposes.
2. Check the water leaks in pipes and repair them properly.
3. Reuse the soapy water, after washing clothes for washing courtyard, carpets etc.
4. Use drip irrigation.
5. Rain water harvesting system should be installed in all the houses.
6. Sewage treatment plant may be installed in all industries and institution.
7. Continuous running of water taps should be avoided.
8. Watering of plants should be done in the evening.

Conservation of soil:

1. Grow different type plants i.e trees, herbs and shrubs.
2. In the irrigation process, using strong flow of water should be avoided.
3. Soil erosion can be prevented by sprinkling irrigation.

Conservation of food resources:

1. Cook required amount of food.
2. Don't waste the food; give it to someone before spoiling.
3. Don't store large amount of food grains and protect them from damaging insects.

Conservation of forest:

1. Use non timber product.
2. Plant more trees.
3. Grassing must be controlled
4. Minimize the use of paper and fuel.
5. Avoid the construction of dam, road in the forest areas.

c) Equitable use of resources for sustainable lifestyles.

Equitable Use of Resources for Sustainable Life Style

1. Urbanization has changed the life style of middle class population in developing countries creating more stress on the use of natural resources.
2. It has been estimated that More Developed Countries (MDC) of the world constitute only 22% of world's population but they use 88% of natural resources. These countries use 73% of energy resources and command 85% of income and in turn they contribute very big proportion of pollution.
3. On the other hand less developed countries (LDCs) have moderate industrial growth and constitute 78% of world's population and use only 12% of natural resources, 27% of energy and have only 15% of global income.
4. In this age of development the rich have gone richer and the poor is becoming poorer. This has lead to unsustainable growth. There is an increasing global concern about the management of natural resources.
5. The solution to this problem is to have more equitable distribution of resources and income. Two major causes of un sustainability are over population in poor countries and over consumption of resources by rich countries. A global consensus has to be reached for balanced distribution of natural resources.
6. For equitable use of natural resources more developed countries/rich people have to lower down their level of consumption to bare minimum so that these resources can be shared by poor people to satisfy their needs.

Expected Questions

1. Short Answer Type Question:

a) What are natural sources? Classify them.
b) What are renewable resources?
c) Define non-renewable resources.
d) What are the problems associated with the exploitation of natural resources?
e) What is the function of forest resources?
f) What is the various location of water?
g) What are mineral resources? How these are exploited?
h) What are the food resources? Discuss the food problems.
i) Give two examples of renewable resources.
j) Mention the causes of deforestation.
k) How forest resources are exploited?
l) What are the primary natural resources?
m) What is the role of Dam in forest?

2. Long Answer type Question:

a. Describe various kinds of natural resources. Distinguish between renewable and non-renewable natural resources, and conventional and non-conventional natural resources.
b. What is meant by equitable use of resources for sustainable use? How can be natural resources be prevented from being deleted? What are the measures to conserve non-renewable natural resources?
c. Discuss the role of individual in conservation of natural resources.
d. Discuss how the land resources are degraded.
e. Describe the use and over-utilization of surface and ground water.
f. What are the food problems? How these are solved?
g. Discuss the use and over exploitation of forest resources.
h. Describe the use of alternative energy sources
i. Write short notes on
   i. Dam's benefits and problems.
   ii. Soil erosion and desertification.
j. Discuss the environmental effects of extracting and using mineral resources with case studies.

ECOSYSTEMS

UNIT - 3

ECOLOGY:

- The term was first coined by Hons Reiter and Haekelin1869.
- The term ecology (Okekologie) is originated from two Greek words Oikos (eco) – means "house" (or) place of living and "ology" means "the science of (or) the study of. Hence, ecology is the branch of science that deals with the study of the pattern of relations between the organism and their environment.

(OR)

- Ecology is the study of interactions among organisms (or) group of organisms with their environment.

(OR)

- Ecology is the study of ecosystems.
- Concept of an eco-system.

ECO SYSTEM:

- In 1935, the British ecologist A.G.Tansleycoined the term "ecosystem".
- The term **"eco system"** is made up of two Greek words. "Eco" means ecological sphere (or) house (or) place of living (or) surroundings (or) Environment, w here living organism does exist while **"system"** means "group of organisms joined in regular and interdependent manner. Hence,
- A group of organisms interacting among them and with environment is known as ecosystem.

(OR)

- A system of interaction of organisms with their surroundings (i.e., environment) is called as "ecosystem".

Examples: Pond, lake, ocean, forest and desert…. etc. are some of the examples of the ecosystems.

- Structure and function of an eco-system.

CHARACTERISTICS OF ECO SYSTEM

- Eco system is the basic functional unit of ecology.
- It contains both biotic and a biotic components.
- The function of ecosystem is related to the cycling of matter (materials) and flow of energy.
- The amount of energy needed to maintain an ecosystem depends on its structure.
- Ecosystem passes from a less complex state to more complex state, which is called as **"Ecological succession".**

CLASSIFICATION OF ECOSYSTEM:

The ecosystem can be generally classified into two types:

1. Natural Ecosystem 2. Artificial Ecosystem

1. NATURALECOSYSTEM:

- A natural ecosystem is developed and governed by nature.
- These are capable of operating and maintaining themselves without any major interference by man.

The following are the two types of natural ecosystem based on their habitat.

- Terrestrial Ecosystem.
- Aquatic Ecosystem.

1) Terrestrial Ecosystem:

- This ecosystem is related to land.

Examples: Grassland ecosystem, Forest ecosystem, and Desert ecosystem etc.

2) Aquatic Ecosystem:

- This ecosystem is related to water, it is further sub divided into two types based on salt content.

i. Fresh Water Ecosystem:

- Running Water Ecosystems

Examples: Rivers, streams (small narrow rivers)

- Standing Water Ecosystems Examples: Pond, Lake Well etc.

ii. Marine Ecosystem:

Examples: seas and sea shores <land along the edges of sea>

2. MAN MADE (OR) ARTIFICIAL ECOSYSTEM:

- An artificial ecosystem is created and maintained by man for his different needs.

Examples: Reservoirs, Artificial lakes and gardens, etc.

STRUCTURE (OR) COMPONENTS OF ANECOSYSTEM:

- The term structure refers to various components. So, the structure of an ecosystem explains the relationship between the Abiotic (non-living) and the biotic (living) components.
- Each and every ecosystem has two major components are:

  i. Biotic (living) components.

  ii. Abiotic (Non-living) components.

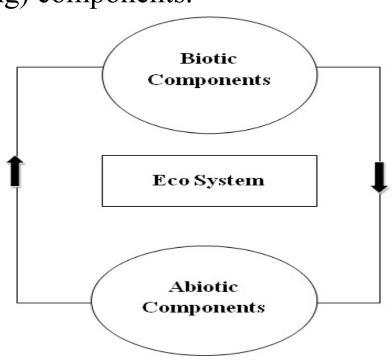


Fig: Components Of Ecosystem

1. Biotic Components: The living component of an ecosystem is called **"Biotic component".**

Examples: Plants (Producers) , Animals (Consumers) and Micro Organisms (Decomposers)

- The biotic components of an ecosystem are classified into three types based on how they get their food.

A. Producers (Autotrophs) :Plants

B. Consumers (Heterotrophs) : Animals

C. Decomposers (Saprotrophs): Micro Organisms.

A. Producers (or) Autotrophs (Auto=self, troph=feeder)

- Self-food producing organisms are known as Autotrophs. Examples: All green plants and trees.
- Producers synthesize their food themselves through photosynthesis. Hence they are also called **"Photo Autotrophs".** (photo =light)

B. Consumers (or) Heterotrophs (Hetero = other, troph=feeder:

- Consumers are organisms, which cannot prepare their own food and depend directly (or) indirectly on the producers.

Examples: Plant Eating Species: Insects, rabbit, goat, deer, cow etc.

Animals Eating Species: Fish, lions, tigers etc.

- Depending upon the food habits the consumers are divided into four types.
  - A. Herbivores (or) Primary Consumers (Plant Eaters)
  - B. Carnivores (or) Secondary Consumers (Meat Eaters)
  - C. Omnivores (or) Tertiary Consumers (With plant & meat eaters)
  - D. Detritivores (dead organism eaters)

A. Herbivores: (Herbi= the green plant &Vorare= to devour)

- Animals that eat only plants are called Herbivores.
- They directly depend on the plants for their food. So they are called Plant eaters.

Examples: Insects, goat, deer, cow, horse etc.

B. Carnivores: (Carne = flesh meat &Vorare= todevour)

- Animals that eat other animals are called carnivores.
- They directly depend on the herbivores for their food.

Examples: Frog, cat, snake &foxes etc.

C. Omnivores: (Omni = whole comes from "ohm" & Vorare = to devour)

- Animals that eat both plants and animals are called omnivores.
- They depend on both herbivores and carnivores for their food.

Examples: humans, tigers, lions, rats and fox etc.

D. Detritivores: (Detritifeeder)

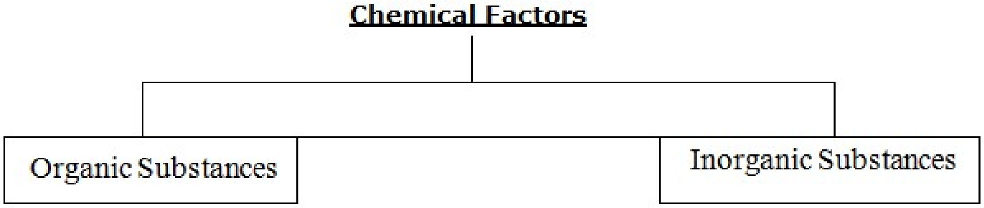
- Animals that eat dead organisms and waste of living are called detritivores.

Examples: beetles, termites, ants, crabs, earthworm etc.

C. Decomposers (or) Saptrotrophs: (Saptro = Rotten, trophos=feeder)

- Decomposers attack the dead bodies of producers and consumers and decompose them into simple compounds. During the decomposition inorganic nutrients are released.
- The organisms which break down the complex compounds into simple products are called decomposers (or) reducers.
- The non-living component of an ecosystem is called "A biotic component"
- These non-living components enter the body of living organism, take part in metabolic activities and then return to the environment. The A biotic component of the ecosystems divided into three portions.
- Climate factors: Solar radiation, temperature, wind, water current, rain fall etc.
- Physical factors: light, fire, soil, air etc.
- Chemical factors: Organic and Inorganic substances.

Chemical Factors



FUNCTION OF AN ECOSYSTEM:

- The function of an ecosystem is related to the cycling of materials (matter) and flow of energy.

Types of functions:

- Functions of an ecosystem are of three types:

1. Primary Function: The producers (plants) can make their food themselves through photosynthesis. This process is called primary function of ecosystem.

Examples: All green plants and trees.

2. Secondary Function: The consumers (animals and humans) cannot make their own food. They are always depending upon the producers for their energy. This is called secondary function of ecosystem.

3. Tertiary Function: Decomposers attack the dead bodies of consumers and producers and decompose the min to simpler compounds. During the decomposition inorganic nutrients are released.

Examples: Micro organisms like bacteria and fungi, etc.

Food chains, food webs and ecological pyramids.

The functioning of an ecosystem may be understood by studying the following terms:

A. Food chains
B. Food webs
C. Food pyramids (or) Energy pyramids
D. Energy and material flow.

A. Food Chain:

- Anything which we eat to live is called food.
- Food contains energy.
- Food can be transferred from one organism to the other.
- The process of transfer of food (energy) from one organism to a series of organisms is called as **"Food Chain".**
- A food chain always starts with a plant life and end with animal life. Thus a food chain is a picture (or) model that shows the flow of energy from Autotrophs (producers) to series of organisms in an environment, as shown in the following figure.

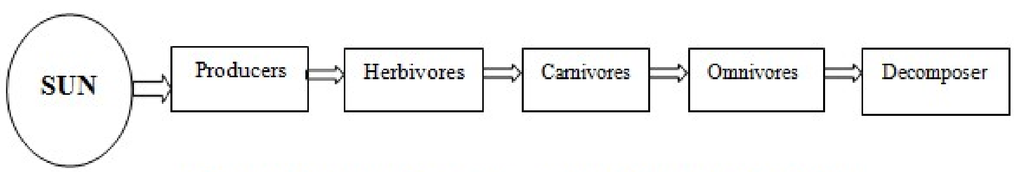


Figure: Schematic representation of food chain.

- In fact all the food chains start with the sun. The sun provides energy for plants.
- The producers (plants) can make their food themselves with the help of the sunlight, chlorophyll, water and air. The consumers, including animals and humans, cannot make their own food. They are always depending upon the producers for their energy.
- Decomposers are the micro-organisms that break down the dead animals and plants and release nutrients that become part of the soil, which are re-used by new plants, back to the starting point of the food chain.

Types of food chain:

Three basic types of food **chains** are found in a typical eco system. They are:

1. Grazing food chains. 2. Detritus food chains. 3. Parasitic food chains.

1. Grazing food chains:

- Grazing food chain starts with green plants (producers) and goes to decomposer food chain (or) detritus food chain through herbivores and carnivores.
- It has two types:

a. Terrestrial food chain and b. Aquatic food chain

Terrestrial food chain: Food chain on land is called terrestrial food chain.

Example: Grassland food chain, Forest land food chain & Desert land food chain.

Grass land food chain

Grasses → Grasshoppers → Frog → Snake → Eagles

Forest food chain

Green plants → Deer → Tiger (or) lion

Aquatic food chain: This food chain is slightly different from terrestrial food chain. It is seen in aquatic (water) eco system. Food chain in water is called "Aquatic food chain".

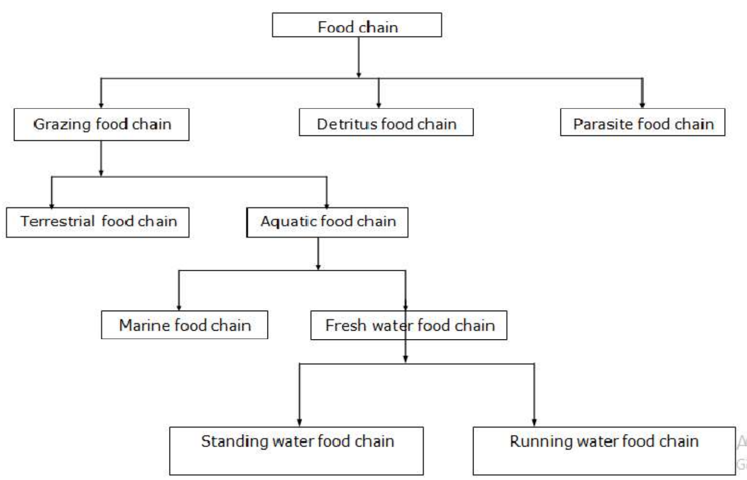
Example: Marine food chain, Ocean, Fresh water food chain, Pond, lake, streams etc.

Food chain in a pond

Phytoplankton → Zoo Plankton → Small fish → Large fish → Man

Marine Food chain:

Sea Weeds → Small fish → Large fish → Sharks and other animals



Detritus' food chain: Detritus food chain starts with dead organic matter (plants and animals) and goes to decomposer through consumers. Detritus food chains, independent of solar energy, but they depend on influx of dead organic matter.

Example: Dead Plants →Soil mitts Algae→ Crabs Small fish→ large fish

Parasitic food chain: Parasitic food chain operates in many ecosystems. In this food chain either consumer (or) producer is parasitized and the food passes to smaller organisms. A parasitic food chain involves host parasite hyper parasites' links.

Example: Trees→ Fruit eating birds→ Lice &Bugs→ Bacteria→ Fungi

Food Web:

- Web means "network" such as spider's web, World Wide Web (WWW) etc.
- So, food web is a network of food chains.
- In a food web many food chains are inter connected, where different types of organisms are connected at different tropic levels, so that there are a number of options of eating and being eaten at each tropic level. Thus, there is a inter connecting of various food chains are called food webs and as shown in following figure.

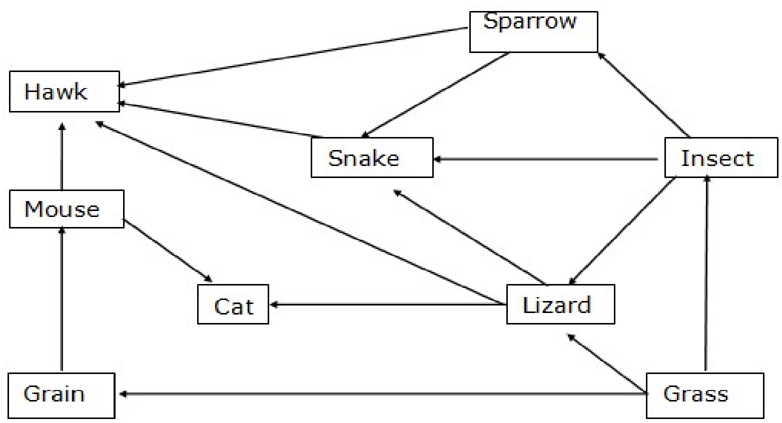


Fig. Food Web

This food web shows many linear food chains <as shown in figure>. These linear food chains are inter connected with other food chains operating in the eco system to forma food web. The grazing food chains are as follows:

- Grains → Mouse → Cat
- Grains → Mouse → Hawk
- Grains → Mouse → Snake → Hawk
- Grains → Insect → Sparrow → Hawk
- Grass → Insect → Lizard → Snake → Hawk
- Grass → Insect → Sparrow → Snake → Hawk

The above food web is a simple one. Much more complex food webs do exist in nature.

Ecological Pyramids:

- The concept of ecological pyramids was first developed by British ecologist Charles Elton in1927.
- Ecological pyramids are the diagrammatic representation of tropic structures in which the tropic levels (i.e., tiers) are depicted in successive stages.
- An ecological pyramid is shown in the following figure.
- In ecological pyramids, tropic levels are shown in the following manner:

  i. The producers represent first tropic level in the ecological pyramid.

  ii. The herbivores (or) primary consumers represent second tropic level in the ecological pyramid.

iii. The carnivores (or) secondary consumers represent third tropic level in the Ecological pyramid.
iv. The omnivores (or) tertiary consumers represent fourth tropic level in the ecological pyramid.

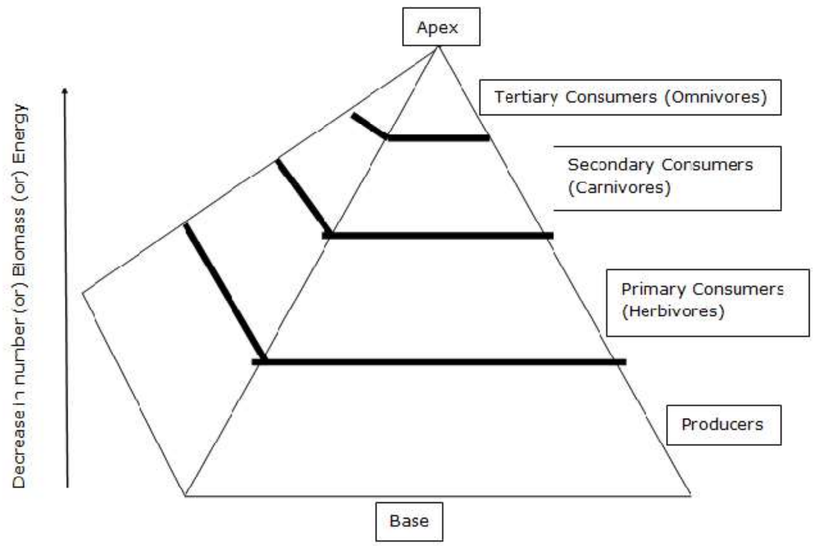


Figure: Formation of an Ecological Pyramid

• On the basis of the number of organisms, the biomass of organisms and energy flow in organist population. Three types of ecological pyramids are:
1. Pyramid of numbers.
2. Pyramid of biomass
3. Pyramid of energy.

Pyramid of numbers:

- It shows the number of individual organisms present in each tropic level.
- It is expressed in numbers per unit area.
- Depending upon the type of ecosystem, we have three types of pyramid of numbers.
  a) Upright pyramid of numbers.
  b) Partly upright pyramid of numbers.
  c) Inverted pyramid of numbers.

Upright Pyramid of numbers:

- The number of individual organisms gradually decreases from lower tropic level to higher tropic level is called ***"upright pyramid of numbers"***.
  Example: A grassland ecosystem and a pond ecosystem show an upright pyramid of numbers.
- The producers in the grass lands are grasses, which are small in size and large in numbers. So, producers occupy lower tropic level (1sttropiclevel).
- The primary consumers (herbivores) are rats, which occupy the II tropic level. Since the numbers of rats are lower when compared to the grasses, the size of which is lower.
- The secondary consumers (carnivores) are snakes, which occupy the III tropic level. Since the numbers of snakes are lower when compared to the rats, the size of which is lower.

- The tertiary consumers (omnivores) are eagles, which occupy the IV tropic level. The number and size of the last tropic level is lowest <as shown in figure>.
- Similarly, in the case of pond ecosystem, producers, herbivores and carnivores are decreases from lower tropic level to the higher tropic level. Thus, these pyramids are upright.
- Therefore, the numbers of individual organisms permit area, decreases from lower tropic level to higher tropic level as shown in figure.

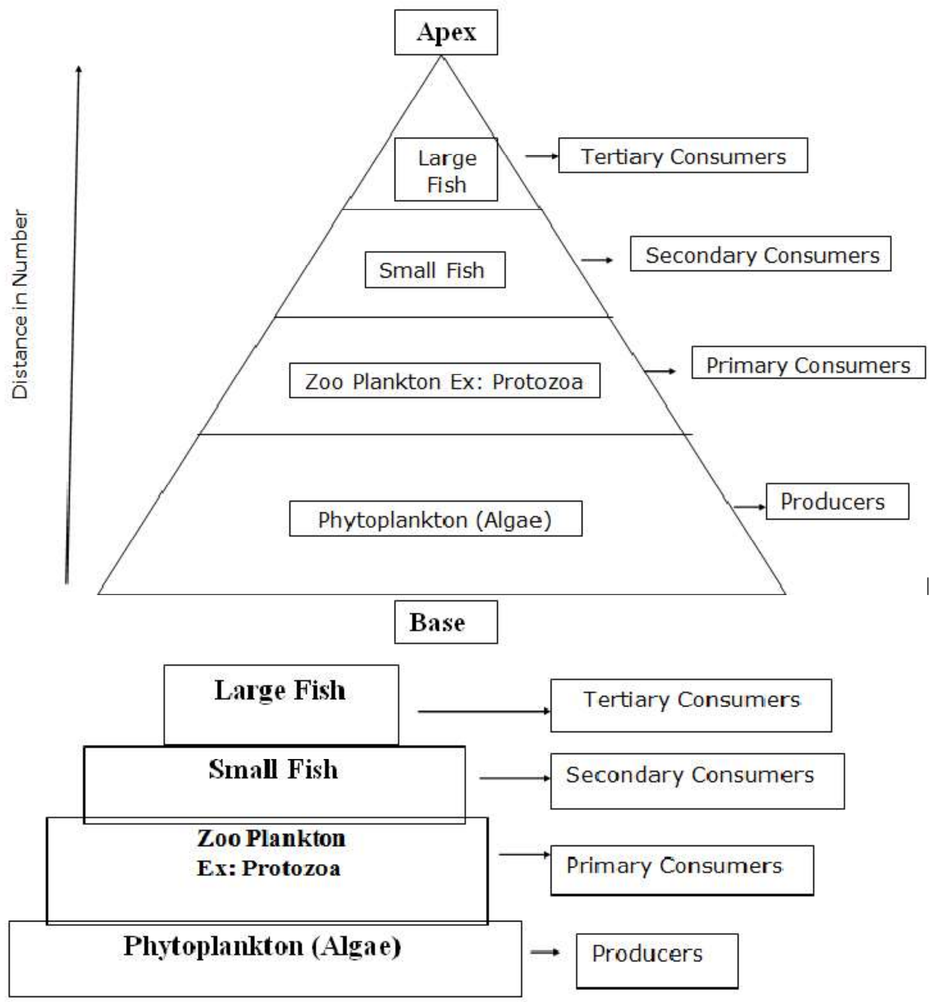


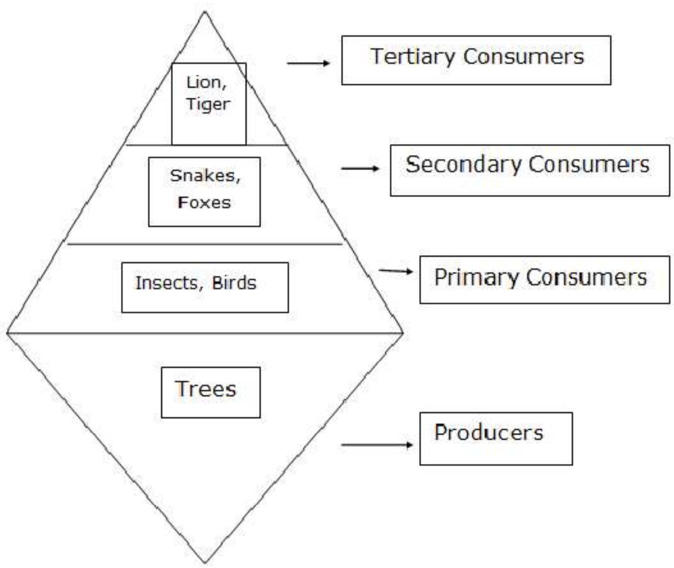
Figure: Pyramid of numbers in an aquatic (pond) ecosystem

Partially Upright Pyramid of Numbers:

- A forest eco system is an example of partially upright pyramid.
- In a forest eco system, big trees are the producers, which are less number. So, these producers occupy the lower tropic level which is narrow base.
- The primary consumers (herbivores) are birds, insects, which occupy the II tropic level. Since the number of birds, insects and other species are higher when compared to the trees, the size of which is broader.
- The secondary consumers (Carnivores) are fox, snakes, lizards, which occupy the third tropic level. Since

the number of fox, snakes are lower when compared to the birds, insects the size of which is lower.

- The tertiary consumers (omnivores) are lion, tiger, which occupy the IV tropic level. Since the number of lion, tiger are lower when compared to the fox and snakes the size of which is very (or) narrow lower. So the pyramid is narrow on both sides and broader in the middle and hence it is called partially upright of number as shown in figure.



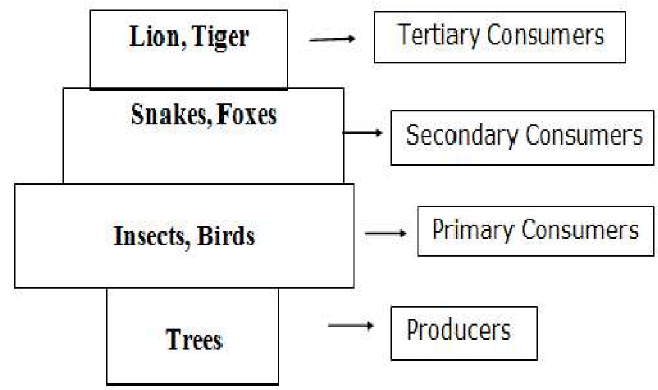
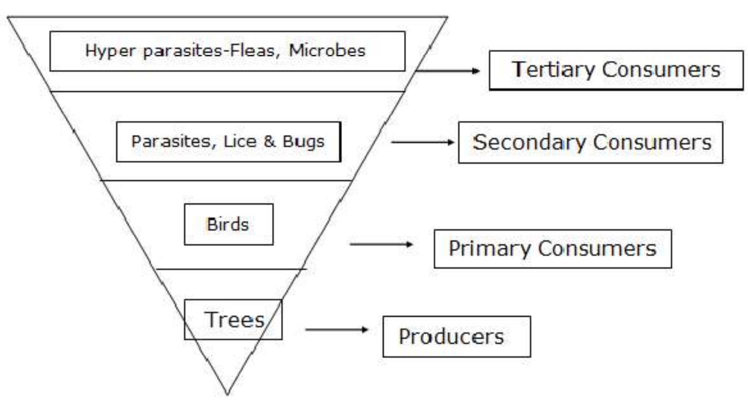


Figure: Pyramid of numbers in the forest ecosystem

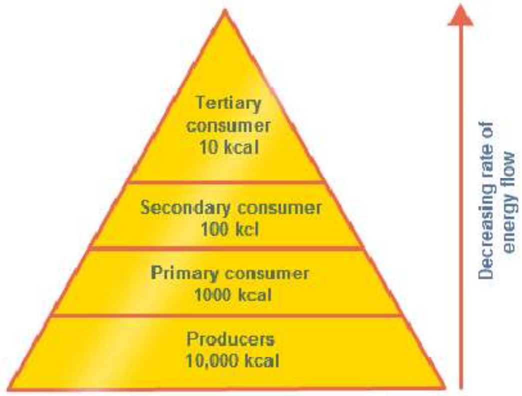
Inverted Pyramid of Numbers:

The number of individual organisms gradually increases from lower tropic level to higher tropic level, is known as "**inverted pyramid of numbers".**

Example: Parasitic food chain shows as inverted pyramid of number as shown in the following figure.



The pyramid of energy or the energy pyramid describes the overall nature of the ecosystem. During the flow of energy from organism to other, there is considerable loss of energy in the form of heat. The primary producers like the Autotrophs there is more amount of energy available. The least energy is available in the tertiary consumers. Thus, shorter food chain has more amount of energy available even at the highest trophic level.



Pyramid of Energy

Forest ecosystems:

Forest ecosystems are areas of the landscape that are dominated by trees and consist of biologically integrated communities of plants, animals and microbes, together with the local soils (substrates) and atmospheres (climates) with which they interact.

Structure of Forest Ecosystems:

Different organisms exist within the forest layers. These organisms interact with each other and their surroundings. Each organism has a role or niche in sustaining the ecosystem.

Some provide food for other organisms; others provide shelter or control populations through predation:

Producers:

- In a forest ecosystem, trees and other plants get their energy from sunlight. Plants produce their own food, in the form of carbohydrates.
- Plants are, therefore, called the primary producers, since they produce the basic foodstuffs for other organisms within food chains and food webs.
- Photosynthesis is the chemical reaction that allows plants to produce their own food.

Consumers:

- Animals cannot produce their own food. They must consume food sources for energy they need to survive.
- All animals, including mammals, insects, and birds, are called consumers.
- Consumers rely on plants and other animals as a food source.
- **Primary consumers** only eat plants and are referred to as herbivores.
- **Secondary consumers** are referred to as carnivores and feed on herbivores.
- **Tertiary consumers** are carnivores that feed on other carnivores.
- **Omnivores** eat both plant and animal matter.

Decomposers:

- Leaves, needles, old branches, dead plants and animals are decomposed by worms, microbes, fungi, ants, and other bugs.
- Decomposers break these items down into their smallest primary elements to be used again. Decomposers

are important in that they sustain the nutrient cycle of ecosystems.

- Humans are part of Forest Ecosystem:
- Humans are consumers. We get food and materials from forests. Because of this, we are a part of the forest ecosystem. Human consumption alters forest ecosystems. Human intervention may be necessary to sustain forest communities under the increased pressure of human use.

GRASSLAND ECOSYSTEM:

Dominated by grass – few shrubs and trees are also found – rainfall average but erratic – overgrazing leads to desertification.

Three types – depending on the climate

- **Tropical grass lands –** found near the boarders of tropical rain forests. Eg. Savannas in Africa. Animals – Zebra, giraffes etc. – fires are common in dry seasons – termite mounds produce methane – leads to fire – high in photosynthesis – deliberate burning leads to release of high CO2 – global warming.
- **Temperate grasslands –** flat and gentle slopes of hills. Very cold winter and very hot summer - dry summer fires do not allow shrubs and trees to grow – soil is quite fertile – cleaned for agriculture.
- **Polar grasslands –** found in arctic polar region – organism – arctic wolf, fox, etc. – A thick layer of ice remains frozen under the soil surface throughout the year – known as permafrost – summer insects and birds appear.

Components:

Structural Components:

Abiotic: soil pH, nutrients, soil moisture, temp, climatic conditions, etc. Biotic: grass, caterpillar, butterfly, worms, insects, birds, etc.

- **AQUATICECOSYSTEM**

Definition: Deals with water bodies and biotic communities present in them-Classified as fresh water and marine ecosystems. Fresh water systems are classified as lentic and lotic ecosystems.

Types:

a) Pond ecosystem: Small fresh water ecosystem – seasonal in nature– organisms: algae, aquatic plants, insects, fishes etc. Ponds are very often exposed to anthropogenic pressure like cloth washing, bathing, cattle bathing, swimming etc.

b) **Lake Ecosystem:** Big fresh water ecosystem – Zonation or stratification, especially during summer is a common one.

Top layer – shallow, warm, prone to anthropogenic activities – Littoral zone

Second layer – enough sunlight, high primary productivity – Limnetic zone

Third layer – very poor or no sunlight – Profundal zone

Eg. Dal Lake in Srinagar, Naini Lake in Nainital

Organisms:

- Planktons – phytoplankton eg. Algae – zooplankton eg.Rotifers
- Nektons – that swims in water eg.Fishes
- Neustons– that float on the surface of water Benthos – that attached to sediments eg. Snails

Types of lakes: Many types

- Oligotrophic lakes – with less nutrient content

- Eutrophic lakes– with very high nutrient content due to fertilizer contamination
- Desert salt lakes – that contains high saline water due to over evaporation
- Volcanic lakes – formed by water emitted from magma due to volcanic eruptions
- Dystrophic lakes – that contains highly acidic water (lowpH)
- Endemic lakes – lakes that contain many endemic species etc.

Streams: fresh water ecosystem where water current plays a major role. Oxygen and nutrient content are uniform. Stream organisms have to face extreme difference in climatic conditions but they do not suffer from oxygen deficiency as pond and lake organisms. This is because large surface area of running water provides more oxygen supply. The animals have very narrow range of tolerance towards oxygen deficiency. Thus stream are worst victims of industrial pollution.

River ecosystem: large streams flowing from mountain highlands are rivers.

- ❖ Three phases:

a) Mountain highlands – rushing down water fall of water – large quantity of dissolved oxygen – plants attached to rocks and fishes that require more oxygen are found.

b) Second phase – gentle slopes of hills – warmer – supports the growth of plants and fishes that require less oxygen are seen.

c) Third phase: river shapes the land – lots of silts, nutrients are brought – deposited in plains and delta – very rich in biodiversity.

Oceans: Gigantic reservoirs of water covering >70% of earth surface – 2,50,000 species – huge variety of sea products, drugs etc. – provide Fe, Mg, oils, natural gas, sand etc. – major sinks of carbon dioxide – regulate biochemical cycles.

❖ Two zones:

a) Coastal zone – warm, nutrient rich, shallow – high sunlight – high primary productivity.

b) Open sea – away from continental shelf – vertically divided in to 3zones.

- Euphotic zone – abundant sunlight
- Bathyal zone – dim sunlight
- Abyssal zone – dark zone – world's largest ecological unit.

❖ **Estuary:** coastal area where river meet ocean – strongly affected by tidal actions – very rich in nutrients – very rich in biodiversity also – organisms are highly tolerant – many species are endemic – high food productivity – however to be protected from pollution.

Characteristics:

Structural Components:

Abiotic: pH, nutrients, D.O, temp, climatic conditions, etc.

Biotic: Phytoplankton, fishes, snails insects, birds, etc.

Expected Questions

1. Short Answer Type Question:

a) What do you mean by ecosystem? Classify the ecosystems.
b) What is ecological succession? Discuss the types of succession.
c) Give the relation between food chain and food web.
d) Discuss different types of consumers.
e) What are dreetrivos? Classify them and give examples.
f) Distinguish between pond ecosystem and Lake Ecosystem.
g) Name different types of ecosystem with examples.
h) What are the characteristic features of desert ecosystem?
i) Define ecology.
j) What is ecological pyramid?

2. Long Answer type Question:

a) Illustrate the energy flow diagram.
b) Discuss the major ecological organization levels.
c) Explain the dimensions of environment.
d) Explain structure of the environment.
e) Earth is an ecosystem. Explain.
f) What is ecological succession? Discuss its types and various phases.
g) Write short notes on
    i. Food chain and food web.
    ii. Desert ecosystem.
    iii. Pond ecosystem.
    iv. Estuaries ecosystem.
h) Describe forest ecosystem with its structure and function.
i) Enumerate the grassland ecosystem and its structure and function.
j) Explain ecological pyramid with suitable diagram.
k) Discuss general characteristics of ecosystem.
l) Differentiate between primary succession and secondary succession.

UNIT 4

Biodiversity and its Conservation

Introduction-Definition: genetics, species and ecosystem diversity.

- Biodiversity is the abbreviated word for ―biological diversity (bio -life or living organisms, diversity-variety). Thus biodiversity is the total variety of life on our planet, the total number of races, varieties and species. The sum of total of various types of microbes, plants and animals (producers, consumers and decomposers) in a system.
- Biomes can be considered life zones, environment with similar climatic, topographic and soil conditions and roughly comparable biological communities (Eg. Grassland, forest). The biomes shelter an astounding variety of living organisms (from driest desert to dripping rain forest, from highest mountain to deepest ocean trenches, life occurs in a marvelous spectrum of size, shape, colour and inter relationship). The variety of living organisms, the biodiversity, makes the world beautiful.
- There are 1.4 million species known presently. But based on new discoveries, by research expeditions, mainly in tropics, taxonomists estimate there are between 3-50 million different species may be alive today. Insects makeup more than one half of all known species and may comprise more than 90% of all species on earth.

LEVELS OFBIODIVERSITY

❖ The concept of biodiversity may be analyzed in 3 different levels. They are

1. Ecosystem Diversity
2. Species Diversity
3. Genetic Diversity

1. Community or Ecosystem Diversity -:

- A set of biotic components (plants, animals and microorganisms) and abiotic components (soil, air, water, etc.) interacting with each other is known as an ecosystem.
- Ecosystem or ecological diversity means the richness and complexity of a biological community, including tropic levels, ecological processes (which capture energy), food webs and material recycling.
- The diversity at an ecological level or habitat level is known as ecosystem diversity.

Ex: River ecosystem- Rivers include fish, aquatic insects, mussels and a variety of plants that have adapted.

- Ecosystem diversity is the aggregate of different environmental types in a region.
- It explains the interaction between living organisms and physical environment in an ecosystem.

2. Species diversity–:

- A discrete group of organisms of the same kind is known as species.
- Species diversity is the diversity between different species.
- The sum of varieties of all living organisms at the species level is known as species diversity.
- Species diversity describes the number of kinds of organisms within individual communities or ecosystems.
- The biotic component is composed of a large number of species of plants, animals and microorganisms

which interact with each other and with the abiotic component of the environment.

Ex: The total number of species living on earth is approximately more than 2 million. However, only around 1.5 million are found and assigned scientific names.

Plant species: Apple, Mango, Wheat, Grapes, Rice etc.

Animal species: Lion, Tiger, Elephant, Deer etc.

3. Genetic Diversity–:

- A species with different genetic characteristics is known as a sub-species or "genera".
- Genetic diversity is a measure of the variety of versions of same gene within individual species.
- With in individual species, there are varieties that are slightly different from one other. These differences are due to differences in the combination of genes.
- Genes are the basic units of hereditary information transmitted from one generation to the other.

Ex: (i) Rice varieties - All rice varieties belong to the species ***"oryzasativa"***.

However there are thousands of rice varieties that show variation at the genetic level in the form of different size, shape, colour and nutrient content.

(ii) Teak wood varieties: The various teak wood varieties available are Indian teak, Burma teak, Malaysian teak etc.

FUNCTIONS OF BIODIVERSITY:-

Two main functions of biodiversity are

1. It is the source on which the entire human species depends on for food, fibre, shelter, fuel and medicine.
2. It depends on biosphere which in turn leads to stability in climate, water, soil, air and overall health of biosphere.

Biogeographically classification of India:-

• India has different climate and topography in different parts and hence is termed as a mega diversity country.

• India occupies **10th place among plant rich countries of the world.**

• It is essential to acquire knowledge about the distribution and environmental interaction of flora and fauna of India.

• Bio-geographers have classified India into ten bio-geographic zones with each zone having characteristic climate, soil and biodiversity.

• These zones are described below:

Trans-Himalayas: The trans-Himalayas are an extension to the Tibetan plateau. This region harbors the high-altitude cold desert in ladakh (Jammu and Kashmir) and Lahaul Spiti (Himachal Pradesh). It accounts for 5.7% of the country's landmass.

Himalayas: The Himalayas are the northern boundaries of India. The entire mountain chain is running from Kashmir in the North-west to Assam in the north-east. The Himalayas comprise of a diverse range of biotic provinces and biomes. The Himalaya as cover 7.2% of the country's landmass

Desert: The extremely dry area west of the Aravalli hill range comprises both the salty desert of Gujarat and the sandy desert of Rajasthan. Deserts occupy around 6.9% of the country's land mass.

The kinds of deserts found in India are:

1. The desert of western Rajasthan
2. The desert of Gujarat
3. The high-altitude cold desert of Jammu &Kashmir and Himachal Pradesh.
4. The Indian deserts have more diversified fauna.

Lakshadweep islands are included in this but the area of these islands is negligible.

Value of biodiversity: consumptive use, productive use, social ethical, aesthetic and optional values.

VALUE OF BIODIVERSITY:-

Definition and estimation of the value of biodiversity is not easy. The value of biodiversity is classified into:

1. Direct Value and
2. Indirect Value

1. Direct value of biodiversity: It is of two types

a. Consumptive use value and b. Productive use value

a. Consumptive use value:

- The consumptive use value is the value placed on nature's products that are consumed directly, without passing through a market. Some of them are firewood, food, and game meat.
- When direct consumption requires recreation, as in sport fishing and game viewing, the consumptive value is the whole recreational experience. Consumptive value seldom appears in national income accounts, but could be easily included in measures such as GDP. It is valued from the cost if resource was sold at market value, rather than being consumed.
- High consumptive use values on resources may lead to the following problems:

- Over-exploitation of wildlife in developing countries
- Loss of traditional control son hunting and
- Loss of wildlife populations at productive levels.
- Consumptive use value benefits the communities closest to the resource if harvested sustainably and manage deficiently.

b. Productive use value:

- Productive use value refers to products that are commercially harvested (sold in a market).
- Its value is estimated at the production end rather than retail end by adding an inflated cost to the finished product.
- Productive use value is often the only value of biological resource reflected in national income accounts and may have a major impact on the national economy.
- Timber, fish, honey, construction materials, mushrooms, fruits, medicinal plants and game meat sold in a market have productive use value.

2. Indirect Value Of biodiversity:

- Indirect values provide economic benefits without being harvested and do not appear in GDP. However, they are crucial to other natural products which influence the GDP.
- These values involve functions performed by biodiversity which are not of any use. Ex: Ecological Processes etc.
- Direct values are often derived from indirect values because plants and animals are supported by the services provided by their environments.
- Many classes of plant and animal species are consumed by tribal and non-tribal communities.

Ex:

1. Ecological functions
2. Flood and storm protection
5. Nutrient cycles
6. Photosynthesis

3. Waste as simulation
4. Microclimatic functions
7. Carbon stores
8. Soil protection etc.

Indirect value of biodiversity is of the following types:

1) Non-consumptive use value
2) Optional value
3) Existence or ethical value and
4) Information value

1) Non-consumptive use value:

- This indirect value deals with nature's functions and services.
- It includes photosynthesis of plants which provides support system for other species by maintaining water cycle, regulating climate, production and protection of the soil, absorption and breakdown of pollutants, recreational, aesthetic, socio-cultural, scientific, educational, spiritual and historic values of natural environments.
- Recreational value is important with regard to tourism and helps the national GDP.

2) Optional value:

- This refers to the potential of biodiversity that is currently known and needs to be explored.
- This refers to the idea that there may be several existing species that may prove to be important in future and their usefulness needs to be studied with reference to a specific problem currently plaguing the society.
- **Ex:** The growing biotechnology field is searching for a cure for diseases like cancer and AIDS.
- Medicinal plants and herbs play a very important role in the economic growth of our country.

3) Existence value:

- This is the value gained from continuous knowledge of existence. Also, this is the value that people are willing to pay to keep a species/ community /ecosystem from going extinct.

Examples of this are high amounts being spent for animals like pandas, whales, lions etc.

- Our rich heritage teaches us to worship plants, animals, rivers and mountains.

Examples being the Ganga River, trees like Banyan and Peepal and plants like the Vambu, Tulsi and Vengai are worshipped.

4) Information value: This relates to the educational, scientific and aesthetic and tourism values of biodiversity in an ecosystem

Aesthetic Values: Beautiful plants and animals inspire us to protect biodiversity. The most important aesthetic value of biodiversity is eco-tourism.

Ex: 1. People from distant places spend time and money to visit areas where they can enjoy aesthetic value of biodiversity. This is called eco-tourism.

2. The pleasant music of wild birds, beautifully coloured butterflies, colour of peacocks and colour of flowers are very important for their aesthetic value.

Threats to biodiversity: Habitats loss, poaching of wild life, man wild life conflicts.

THREATS TO BIODIVERSITY

- Any disturbance in a natural ecosystem tends to reduce its biodiversity.
- Waste generated due to increase in human population and industrialization spoils the environment and leads to decreased diversity in biological species.
- Any change in the system leads to a major imbalance and threatens the normal ecological cycle.

❖ **Causes for loss of biodiversity are:**

1. Habitat loss
2. Poaching of wild life and
3. Man-wild life conflicts

1. Habitat loss: The loss of populations of interbreeding organisms is caused by habitat loss. Factors influencing habitat loss are:

a. **Deforestation:** Loss of habitat is mainly caused by deforestation activities. Forests and grasslands are cleared for conversion into agriculture lands or settlement areas or developmental projects. Forests and grasslands are natural home to thousands of species which disintegrate due to loss of their natural habitat.

b. **Destruction of wetlands:** Wetlands, estuaries and mangroves are destroyed due to farming, filling and pollution that cause loss of biodiversity

c. **Habitat fragmentation:** When the habitat is divided into small and scattered patches the phenomenon is called habitat fragmentation. This leads to the disappearance of most wildlife.

d. **Raw material:** To produce hybrid seeds, wild plants are used as raw materials leading to extinction of many wild plant species.

e. **Production of drugs:** Pharmaceutical companies collect wild plants for the production of drugs leading to extinction of several medicinal plant species.

f. **Illegal trade:** Illegal trade of wildlife reduces biodiversity leading to habitat loss

g. **Developmental activities:** Construction of dams in forest areas coupled with the discharge of industrial effluents kills birds and other aquatic life.

2. Poaching of wildlife: Poaching refers to killing animals or commercial hunting. It contributes to loss of biodiversity. Poaching can be of two types listed below:

1. **Subsistence poaching:** This refers to killing animals for survival.
2. **Commercial poaching:** This refers to hunting animals in order to sell their products.

Factors influencing poaching:

1. Human population: Increased human population in India has led to pressure on forest resources, leading to degradation of wild life habitats

2. Commercial activities: Although a ban has been imposed internationally on the trade of products of endangered species, there is a continued smuggling of wildlife products. Since trading of such products is highly profitable, poachers continue to hunt endangered animals and smuggle their fur, skin and tusks to other countries. Wildlife products include furs, horns, tusks, live specimens and herbal products. Richest source of biodiversity lies in developing nations in Asia, Africa and Latin America. Advanced countries like Europe, North America, Japan, Taiwan, and Honkong are the major importers of wildlife products.

3. Man-Wildlife Conflicts: Man-wildlife conflicts arise, when wildlife starts causing immense damage and danger to man. Under such conditions it is very difficult for the forest department officials to convince the affected villagers to gain the villagers support for wild life conservation.

Ex: 1. In Sambalpur, Orissa, several people were killed by elephants. In retaliation, the villagers killed and injured several elephants.

2. In Mysore, elephants were killed by farmers in retaliation to the damaged one by elephants to their cotton and sugarcane fields.

3. Villagers sometimes hide explosives in their fields to ward-off animals which explode when the elephants enter the fields

4. Several people were killed when leopards attacked them in Sanjay Gandhi National Park, Mumbai

Factors influencing man-animal conflicts:-

- Shrinking forest cover compels wildlife to move outside the forest
- Human encroachment into forest area induces a man-wildlife conflict

- Injured animals have a tendency to attack man
- Wild animals venture out of the forest area in search of food
- Villager's set-up electric wiring around their fields. This injures animals (Elephants) who suffer pain and get violent.
- Cash compensation paid by the government is not enough.
- Garbage near human settlements or food crops attracts wild animals.

CONSERVATION OF BIODIVERSITY

The following measures should be taken to conserve biodiversity

- Illegal hunting and trade of animals and animal products should be stopped immediately
- People-at-large should boycott purchasing coats, purse or bags made of animal skin
- Bio-diversity laws should be strengthened.
- Adequate crop and cattle compensation schemes must be started
- Solar powered fencing must be provided with electric current proof trenches to prevent animals from entering fields.
- Cropping pattern should be changed near the forest borders
- Adequate food and water should be made available for wild animals within forest zones.
- Development and construction work in and around forest region must be stopped.
- Biodiversity is one of the important tools for sustainable development. The commercial, medical, genetic, aesthetic, and ecological importance of biodiversity emphasizes the need for its conservation.

Factors affecting biodiversity:

- Biodiversity is disturbed by human activity
- Poaching of animals, over-exploitation of natural sources and degradation of habitats affect biodiversity.
- Marine ecosystems are disturbed due to oil spills and discharge of effluents
- Climatic factors like global warming, ozone depletion and acid rain also affect biodiversity

Need for biodiversity

- It provides recreation and tourism
- Drugs, herbs, food and other important raw materials are derived from plants and animals
- It preserves the genetic diversity of plants and animals
- It ensures sustainable utilization of life supporting systems on earth.
- It needs to conservation of essential ecological diversity and life supporting systems
- Loss of biodiversity leads to ecological and environmental deterioration

Types of conservation

There are two types of biodiversity conservation:

1. In-situ conservation and
2. Ex-situ conservation

1. IN-SITU CONSERVATION

In-situ conservation involves protection of flora and fauna within its natural habitat. The natural habitats or ecosystems under in-situ conservation are called "protected areas".

a. Biosphere reserves   b. National parks

c. Wild life sanctuaries   d. Gene sanctuaries

a. Biosphere reserves: They cover large areas (>5000 sq.km.) They are normally used to protect species for a long time. The roles of biosphere reserves are listed below:

- Long-term survival of evolving ecosystem

- Protect endangered species
- Protect maximum number of species and communities
- Serve as site of recreation and tourism
- May also be used for educational and research purposes
- Biosphere reserves function as an open system and changes in land use are not allowed. No tourism and explosive activities are allowed in biosphere reserves.

b. A national park: It is an area dedicated for the conservation of wildlife along with its environment. It covers an area ranging from 100 to 500 sq.km. One or more national parks may exist within a biosphere reserve. A national park issued for enjoyment through tourism, without affecting the environment. It is used to protect, propagate and develop wildlife. Grazing domestic animals inside national parks is prohibited All private rights and forestry activities are prohibited inside a national park.

c. Wildlife sanctuary is an area that is reserved for the conservation of animals only.

- It protects animals only
- It allows operations such as harvesting of timber, collection of forest products, private ownership rights and forestry operations, provided it does not affect animals adversely

d. Gene sanctuary is an area where plants are conserved. Other projects for of the conservation animals are Project Tiger, Girl Lion Project, Crocodile breeding project, project elephant etc.

Advantages of in-situ conservation

- It is cheap and convenient
- Species get adjusted to natural disasters like drought, floods, forest fires etc.

Disadvantages of in-situ conservation

- A large surface area of earth is required to preserve biodiversity
- Maintenance is not proper due to shortage of staff and pollution

2. EX-SITU CONSERVATION

- Ex-situ conservation involves protection of flora and fauna outside their natural habitats.
- This type of conservation is mainly done for conservation of crop varieties and wild relatives of crops.
- Ex-situ conservation involves maintenance and breeding of endangered plant and animal species under controlled conditions
- It identifies those species that are at a high risk of extinction
- It prefers species that are important for man in the near future among the endangered species.

Important centers of ex-situ conservation:

1. Botanical gardens	2. Seed banks	3. Microbial culture collections
4. Tissue and cell cultures	5. Museums	6. Zoological gardens

Methods of ex-situ conservation

National Bureau of Plant Genetic Resources (NPBGR):- It is located in New Delhi and uses the Cryopreservation Technique to preserve agricultural and horticultural crops. Cryopreservation technique involves using liquid nitrogen at -196 C. Varieties of rice, turnip, radish, tomato, onion, carrot, chili, tobacco have been successfully preserved for years using this technique.

National Bureau of Animal Genetic Resources (NPAGR):- It is located in Karnal, Haryana and preserves the semen of domesticated bovine animals.

National Facility for Plant Tissue Culture Repository (NFPTCR): -In this facility, conservation of varieties of crop plants or trees is done using tissue culture. This facility has been created within the NPBGR.

Advantages of Ex-situ conservation

1. Survival of endangered species is increasing due to special care and attention.
2. In captive breeding the animals are assured of food, water, shelter and security thereby have a longer lifespan.
3. It is carried-out in cases of endangered species that do not have any chance of survival in the wild.

Disadvantages of Ex-situ conservation

1. It is an expensive method.
2. Freedom of wildlife is lost.
3. Animals cannot survive in the natural environment.

Expected Questions

1. Short Answer Type Question:

a) What is biodiversity?

b) What are different forms of biodiversity?

c) What are the threats to biodiversity?

d) Define in-situ and ex-situ conservation of biodiversity.

e) Discuss the values of biodiversity at the local and the global levels.

f) Define Hotspots.

g) Which is the largest national park in India?

h) How many biodiversity hot spots are in India?

2. Long Answer type Question:

a) What is meant by biodiversity? Explain the hierarchical levels of biodiversity.

b) What do you mean by value of biodiversity? Explain the consumptive use and productivity use of biodiversity.

c) Describe the uses and importance of biodiversity.

d) What are hotspots of biodiversity? Discuss their significance. Explain the hotspots that extend into India.

e) Discuss the causes that have posed threats to biodiversity.

f) Write short notes on

i. In-situ conservation. ii. Ex-situ conservation. iii. Sustainable conservation.

g) What are endangered and endemic species of India?

h) Explain biodiversity at global, national and local levels.

ENVIRONMENTAL POLLUTION UNIT- 5

INTRODUCTION

POLLUTION may be defined as an undesirable change in the physical, chemical or biological characteristics of air, water and land that may be harmful to human life and other animals, living conditions, industrial processes and cultural assets. Pollution can be natural or manmade.

- The agents that pollute are called pollutants.

POLLUTANTS : Pollutants are by-products of man's action. The important pollutants are summarized below:

- **Deposited matter—** Soot, smoke, tar or dust and domestic wastes.
- **Gases—**CO, nitrogen oxides, sulphur oxides, halogens (chlorine, bromine and iodine).
- **Metals—**Lead, zinc, iron and chromium.
- **Industrial pollutants—**Benzene, ether, acetic acid etc., and cyanide compounds.
- **Agriculture pollutants—**Pesticides, herbicides, fungicides and fertilizers.
- **Photo chemical pollutants—**Ozone, oxides of nitrogen, aldehydes, ethylene, photochemical smog and proxy acetyl nitrate.
- **Radiation pollutants—**Radioactive substances and radioactive fall-outs of the nuclear test.

Classification of Pollutants

Nature of disposal: On the basis of natural disposal, pollutants are of two types:

1. Non-degradable pollutants: These are the pollutants, which degrade at a very slow pace by the natural biological processes. These are inorganic compounds such as salts (chlorides), metallic oxides waste producing materials and materials like, aluminum cans, mercuric salts and even DDT. These continue to accumulate in the environment.

2. Biodegradable pollutants: These include domestic sewage that easily decomposes under natural processes and can be rapidly decomposed by natural/ artificial methods. These cause serious problems when accumulated in large amounts as the pace of deposition exceeds the pace of decomposition of disposal.

Nature of form: On the basis of the form in which they persist after their release into the environment, pollutants can be categorized under two types:

Primary pollutants: These include those substances, which are emitted directly from some identifiable sources. This include-

a. Sulphur compounds: SO2, SO3, H2S produced by the oxidation of fuel.

b. Carbon compounds: Oxides of carbon (CO+CO2) and hydrocarbons.

c. Nitrogen compounds: NO2 andNH3.

d. Halogen compounds: Hydrogen fluoride (HF) and hydrochloric acid (HCL).

e. Particles of different size and substances: These are found suspended in air. The fine particles below the diameter of 100u are more abundant and include particles of metals, carbon, tar, pollen, fungi, bacteria, silicates and others.

Secondary pollutants: The secondary pollutants are produced by the combination of primary emitted pollutants in the atmosphere.

Ex: In bright sunlight, a photochemical reaction occurs between nitrogen oxides; oxygen and waste hydrocarbons from gasoline that forms peroxy-acetyle nitrate (PAN) and ozone (O3), both of them are toxic components of smog and cause smarting eyes and lung damage.

DEFINITION CAUSES, EFFECTS AND CONTROL MEASURES OF POLLUTION

AIRPOLLUTION

Introduction: Air pollution is one such form that refers to the contamination of the air, irrespective of indoors or outside. A physical, biological or chemical alteration to the air in the atmosphere can be termed as pollution. It occurs when any harmful gases, dust, smoke enters into the atmosphere and makes it difficult for plants, animals and humans to survive as the air becomes dirty.

The WHO defines **air pollution** as the presence of materials in the air in such concentration which are harmful to man and his environment. A number of ingredients find their way in the air and these are mostly gases, which rapidly spread over wide areas.

The composition of Air is given below:

- Nitrogen 78%
- Oxygen 21%
- Argon Less than 1%
- Carbon dioxide 0.037%
- Water vapour remaining
- Ozone, Helium and ammonia Trace amount

Causes of Air pollution:

- **Burning of Fossil Fuels:** Sulfur dioxide emitted from the combustion of fossil fuels like coal, petroleum and other factory combustibles is one of the major causes of air pollution. Pollutants emitting from vehicles cause immense amount of pollution. Carbon Monoxide produced by improper or incomplete combustion emitted from vehicles is another major pollutant along with Nitrogen Oxides that is produced from both natural and manmade processes.
- **Agricultural activities:** Ammonia is a very common by product from agriculture related activities and is one of the most hazardous gases in the atmosphere. Use of insecticides, pesticides and fertilizers in agricultural activities emit harmful chemicals into the air and cause water pollution.
- **Exhaust from factories and industries:** Manufacturing industries release large amount of carbon monoxide, hydrocarbons, organic compounds, and chemicals into the air thereby depleting the quality of air. Petroleum refineries also release hydrocarbons and various other chemicals that pollute the air and also cause land pollution.
- **Mining operations:** Mining is a process wherein minerals below the earth are extracted using large equipment. During the process dust and chemicals are released in the air causing massive air pollution.
- **Indoor air pollution:** Household cleaning products, painting supplies emit toxic Chemicals in the air and cause air pollution.
- **Suspended Particulate matter:** Suspended particulate matter popular by its acronym SPM, is another cause of pollution.

Types of Pollutants Air

- **Primarily air pollutants** can be caused by primary sources or secondary sources. The pollutants that are a direct result of the process can be called primary pollutants. A classic example of a primary pollutant would be the sulfur-dioxide emitted from factories

- **Secondary pollutants** are the ones that are caused by the inter mingling and reactions of primary pollutants. Smog created by the interactions of several primary pollutants is known to be as secondary pollutant.

Common air pollutants

- **Carbon Dioxide:** CO2 content of air has increased by 20% during the last century. CO2 causes nausea and headache. Its increase in the air may cause greenhouse effect, rise in the atmospheric temperature. This may melt the polar ice resulting in rise in level of oceans and flooding of coastal regions.
- **Carbon Monoxide:** It is a very poisonous gas and is produced by incomplete combustion of fuel. If inhaled. It combines with hemoglobin and reduce sits oxygen-carrying capacity. This leads to laziness, reduced vision and death.
- **Oxides of Nitrogen:** These include NO and NO2,which are released by automobiles and chemical industries as waste gases and also by burning of materials. These are harmful and lower the oxygen carrying capacity of blood.
- **Oxides of Sulphur:** SO2 and SO3 are produced by burning of coal and petroleum and are harmful to buildings, clothing, plants and animals. High concentration of SO2 causes chlorosis (yellowing of leaves), plasmolysis, damage to mucous membrane and metabolic inhibition. SO2 and SO3 react with water to form Sulphuric and sulphurous acids. These may precipitate as rain or snow producing acid rain or acid precipitation.
- **Photochemical Oxidants:** Formed by the photochemical reactions between primary pollutants, viz. oxides of nitrogen and hydrocarbons. Nitrogen oxides in the presence of sunlight react with un-burnt hydrocarbons to form peroxy acetyl nitrate (PAN), Ozone, aldehydes and some other complex organic compounds in the air.
- **Hydrocarbons:** These are un-burnt discharges from incomplete combustion of fuel in automobiles. These forms PAN with nitrogen oxides, which is highly toxic.
- **Particulate Matter:** Industries and automobiles release fine solid and liquid particles into the air. Fly ash and soot from burning of coal, metal dust containing lead, chromium, nickel, cadmium, zinc and mercury from metallurgical processes; cotton dust from textile mills; and pesticides sprayed on crops are examples of particulate pollutants in the air. These are injurious to respiratory tract.
- **Aerosols:** Aerosols are chemicals released in the air in vapor form. These include fluorocarbon (carbon compound having fluorine) present in emissions from the Jet aero planes. Aerosols deplete the ozone layer. Thinning of ozone layer results in more harmful ultraviolet rays reaching the earth, which are harmful to skin, and can lead to skin cancer also.
- **Radioactive Substances:** These are released by nuclear explosions and explosives. These are extremely harmful for health.
- **Fluorides:** Rocks, soils and. Minerals containing fluorides release an extremely toxic gas called hydrogen fluoride on heating. This gas is highly injurious to livestock and cattle.

Control Measures

The atmosphere has several built-in self-cleaning processes such as dispersion, gravitational settling, flocculation, absorption, rain-washout, etc. to cleanse the atmosphere. However, control of contaminants at their source level is a desirable and effective method through preventive or control technologies.

A. Source control: Some measures that can be adopted in this direction are

1. Using unleaded petrol
2. Using fuels with low sulphur and ash content
3. Encouraging people to use public transport walk or use a cycle as opposed to private vehicles
4. Ensure that houses, schools, restaurants and playgrounds are not located on busy streets
5. Plant trees along busy streets as they remove particulates, carbon dioxide and absorb noise
6. Industries and waste disposal sites should be situated outside the city preferably on the downwind of the city.
7. Catalytic converters should be used to help control emissions of carbon monoxide and hydro carbons

B. Control measures in industrial centers:

1. Emission rates should be restricted to permissible levels by each and every industry
2. Incorporation of air pollution control equipment in design of plant layout must be made mandatory.
3. Continuous monitoring of the atmosphere for pollutants should be carried out to know the emission levels.

Equipment used to control air pollution

Air pollution can be reduced by adopting the following approaches.

1. Ensuring sufficient supply of oxygen to the combustion chamber and adequate temperature so that the combustion is complete thereby eliminating much of the smoke consisting of partly burnt ashes and dust.
2. To use mechanical devices such as scrubbers, cyclones, bag houses and electro- static precipitators in manufacturing processes. The equipment used to remove particulates from the exhaust gases of electric power and industrial plants are shown below. All methods retain hazardous materials that must be disposed safely. Wet scrubber can additionally reduce sulphur dioxide emissions.
3. The air pollutants collected must be carefully disposed. The factory fumes are dealt with chemical treatment.

WATER POLLUTION

Water pollution may be defined as "the alteration in physical, chemical and biological characteristics of water which may cause harmful effects on humans and aquatic life." Pollutants include:

- Sewage
- Industrial effluents and chemicals
- Oil and other wastes
- Chemicals in air dissolve in rain water, fertilizers, pesticides and herbicides leached from land pollute water.

TYPES, EFFECTS AND SOURCES OF WATER POLLUTION

Water pollution is any chemical, biological or physical change in water quality that has a harmful effect on living organisms or makes water unsuitable for desired uses.

Infectious agents

Ex: Bacteria, Viruses, Protozoa, and parasitic worms.

- Human sources Human and animal wastes

Effects: Variety of diseases.

- **Oxygen demanding wastes (Dissolved oxygen):** This degradation consumes dissolved oxygen in water.

Human sources: Sewage, Animal feedlots, paper mills and food processing facilities.

Effects: Large populations of bacteria decomposing these wastes can degrade water quality by depleting water of dissolved oxygen.

- Inorganic chemicals

Ex: Water soluble inorganic chemicals: Acids Compounds of toxic metals such as lead (Pb), arsenic (As) and selenium (Se) Salts such as NaCl in oceans and fluoride (F-) found in some soils.

Human sources: Surface runoff, industrial effluents and household cleansers Effects: Inorganic chemicals can: Make freshwater unusable for drinking and irrigation Cause skin cancer and neck damage nervous system, liver and kidneys Harm fish and other aquatic life Lower crop yields .Accelerate corrosion of metals exposed to such water.

- **Organic chemicals Ex:** Oil, Gasoline, Plastics, Pesticides, Cleaning solvents and Detergents. Human Sources: Industrial effluents, household cleansers and surface runoff from farms.

Effects: Can threaten human health by causing nervous system damage and some cancers. Harm fish and wildlife.

Plant nutrients Ex: Water soluble compounds containing nitrate, Phosphate and Ammonium ions.

Effects: Drinking water with excessive levels of nitrates lower the oxygen carrying capacity of the blood and can kill urban children and infants.

- **Sediment Ex:** Soil, silt, etc. Human Sources: Land erosion

Effects: Causes cloudy water thereby reducing photosynthetic activity Disruption of aquatic food chain Carries pesticides, bacteria and other harmful substances Settles and destroys feeding and spawning grounds of fish Clogs and fills lakes, artificial reservoirs, stream channels and harbours.

- **Radioactive materials:** Iodine Radon Uranium Cesium and Thorium

Human sources: Nuclear power plants, mining and processing of uranium and other ores, nuclear weapon production and natural sources.

Effects: Genetic mutations, birth defects and certain cancers.

- **Heat (Thermal pollution) Ex**: Excessive heat

Human sources: Water cooling of electric power plants and some types of industrial plants.

Effects: Low dissolved oxygen levels thereby making aquatic organisms more vulnerable to disease, parasites and toxic chemicals.

Point and non-point sources of water pollution:

- **Point sources:** These are pollutants that are discharged at specific locations through pipes, ditches or sewers into bodies of surface waters.

Ex: Factories, sewage treatment plants, abandoned underground mines and oil tankers.

- **Non point sources** These pollutants cannot be traced to a single point of discharge. They are large land areas or air-sheds that pollute water by runoff, subsurface flow or deposition from the atmosphere.

Ex: Acid deposition, runoff of chemicals into surface water from croplands, livestock feedlots, logged forests, urban streets, lawns, golf courses and parking lots.

Control measures of water pollution

- Administration of water pollution control should be in the hands of state or central government
- Scientific techniques should be adopted for environmental control of catchment areas of rivers, ponds or

streams

- Industrial plants should be based on recycling operations as it helps prevent disposal of wastes into natural waters but also extraction of products from waste.
- Plants, trees and forests control pollution as they act as natural air conditioners.
- Trees are capable of reducing sulphur dioxide and nitric oxide pollutants and hence more trees should be planted.
- No type of waste (treated, partially treated or untreated) should be discharged into any natural water body. Industries should develop closed loop water supply schemes and domestic sewage must be used for irrigation.
- Qualified and experienced people must be consulted from time to time for effective control of water pollution.
- Public awareness must be initiated regarding adverse effects of water pollution using the media.
- Laws, standards and practices should be established to prevent water pollution and these laws should be modified from time to time based on current requirements and technological advancements.
- Basic and applied research in public health engineering should be encouraged.

SOIL POLLUTION

Soil pollution is defined as, "contamination of soil by human and natural activities which may cause harmful effect on living organisms". Composition of soil is listed below:

COMPONENT %

Organic mineral matter 45%, Organic matter 05%, Soil water 25%, Soil air 25%

TYPES, EFFECTS AND SOURCES OF SOIL POLLUTION

Soil pollution mainly occurs due to the following:

- Industrial wastes
- Urban wastes
- Agricultural practices
- Radioactive pollutants
- Biological agents

Industrial Wastes – Disposal of Industrial wastes is the major problem for soil pollution

Sources: Industrial pollutants are mainly discharged from various origins such as pulp and paper mills, chemical fertilizers, oil refineries, sugar factories, tanneries, textiles, steel, distilleries, fertilizers, pesticides, coal and mineral mining industries, drugs, glass, cement, petroleum and engineering industries etc.

Effect: These pollutants affect and alter the chemical and biological properties of soil. As a result, hazardous chemicals can enter into human food chain from the soil or water, disturb the biochemical process and finally lead to serious effects on living organisms.

Urban Wastes – Urban wastes comprise of both commercial and domestic wastes consisting of dried sludge and sewage. All the urban solid wastes are commonly referred to as refuse.

Constituents of urban refuse: This refuse consists of garbage and rubbish materials like plastics, glasses, metallic cans, fibers, paper, rubbers, street sweepings, fuel residues, leaves, containers, abandoned vehicles and other discarded manufactured products. Urban domestic wastes though disposed of separately from industrial wastes, can still be dangerous. This happens because they are not easily degraded.

Agricultural Practices – Modern agricultural practices pollute the soil to a large extent. With the advancing agro-technology, huge quantities of fertilizers, pesticides, herbicides and weedicides are added to increase the crop yield. Apart from these farm wastes, manure, slurry, debris, soil erosion containing mostly inorganic chemicals are reported to cause soil pollution

Radioactive pollutants - **Radioactive substances** resulting from explosions of nuclear testing laboratories and industries giving rise to nuclear dust radioactive wastes penetrate the soil and accumulate giving rise to land/soil pollution.

Ex: Radio nuclides of Radium, Thorium, Uranium, isotopes of Potassium (K-40) and Carbon (C-14) are commonly found in soil, rock, water and air.

Explosion of hydrogen weapons and cosmic radiations include neutron, proton reactions by which Nitrogen (N-15) produces C-14. This C-14 participates in Carbon metabolism of plants which is then into animals and human beings.

Biological agents – Soil gets a large amount of human, animal and bird excreta which constitute a major source of land pollution by biological agents.

Ex: 1. Heavy application of manures and digested sludge can cause serious damage to plants within a few years

Control measures of soil pollution:

Soil erosion can be controlled by a variety of forestry and farm practices.

Ex: Planting trees on barren slopes

- Contour cultivation and strip cropping may be practiced instead of shifting cultivation
- Terracing and building diversion channels may be undertaken.
- Reducing deforestation and substituting chemical manures by animal wastes also helps arrest soil erosion in the long term.

➢ **Proper dumping of unwanted materials:** Excess wastes by man and animals pose a disposal problem. Open dumping is the most commonly practiced technique. Nowadays, controlled tipping is followed for solid waste disposal. The surface so obtained is used for housing or sports field.

➢ **Production of natural fertilizers:** Bio-pesticides should be used in place of toxic chemical pesticides. Organic fertilizers should be used in place of synthesized chemical fertilizers. Ex: Organic wastes in animal dung may be used to prepare compost manure instead of throwing them wastefully and polluting the soil.

➢ **Proper hygienic condition:** People should be trained regarding sanitary habits.

Ex: Lavatories should be equipped with quick and effective disposal methods.

➢ **Public awareness**: Informal and formal public awareness programs should be imparted to educate people on health hazards by environmental education.

Ex: Mass media, Educational institutions and voluntary agencies can achieve this.

➢ **Recycling and Reuse of wastes**: To minimize soil pollution, the wastes such as paper, plastics, metals, glasses, organics, petroleum products and industrial effluents etc should be recycled and reused.

Ex: Industrial wastes should be properly treated at source. Integrated waste treatment methods should be adopted.

➢ **Ban on Toxic chemicals:** Ban should be imposed on chemicals and pesticides like DDT, BHC, etc. which are fatal to plants and animals. Nuclear explosions and improper disposal of radioactive wastes should be banned.

NOISE POLLUTION

Noise is defined as, "the unwanted, unpleasant or disagreeable sound that causes discomfort to all living beings". Sound intensity is measured in decibels (dB) that is the tenth part of the longest unit Bel. One dB is the faintest sound that a human ear can hear.

TYPES OF NOISE: Environmental noise has been doubling every ten years. Noise is classified as:

A. Industrial Noise B. Transport Noise and C. Neighborhood noise

A. Industrial Noise: It is sound with a high intensity sound caused by industry machines. Sources of such noise pollution are caused by machines from machines in various factories, industries and mills. Noise from mechanical saws and pneumatic drills is unbearable and a nuisance to the public. The Indian Institute of Oto-Rino Laryngology, Chennai reported that increasing industrial pollution damages the hearing ability by at least 20%.Workers in steel industry, who work close to heavy industrial blowers, are exposed to 112dB for eight hours suffer from occupational pollution.

B. Transport Noise: Transport noise mainly consists of traffic noise from road, rail and aircraft. The number of automobiles on roads like motors, scooters, cars, motor cycles, buses, trucks and diesel engine vehicles has increased enormously in the recent past further aggravating the problem of transport noise.
Noise levels in most residential areas in metropolitan cities are hovering around the border line due to increased vehicular noise pollution. This high level of noise pollution leads to deafening in the elderly.

C. Neighborhood Noise: This type of noise includes disturbance from household gadgets and community. Common sources being musical instruments, TV, VCR, Radios, Transistors, Telephones, and loudspeakers etc. Statistically, ever since the industrial revolution, noise in the environment has been doubling every ten years.

Effects of Noise Pollution

1) Noise pollution affects both human and animal health. It leads to:
2) contraction of blood vessels
3) making skin pale
4) Excessive adrenalin in the blood stream which is responsible for high blood pressure.
5) Blaring sounds are known to cause mental distress
6) Heart attacks, neurological problems, birth defects and abortion
7) Muscle contraction leading to nervous breakdown, tension, etc.
8) The adverse reactions are coupled with a change in hormone content of blood, which in-turn increases heart beat, constriction of blood vessels, digestive spams and dilation of the pupil of the eye.
9) Adverse effects health, work efficiency and behavior. Noise pollution may cause damage to the heart, brain, kidneys, liver and may produce emotional disturbance.
10) The most immediate and acute effect of noise is impairment of hearing that diminishes some part of the auditory system. Prolonged exposure to noise of certain frequency pattern leads to chronic damage to the inner ear.

Control Measures:

- **SOURCE CONTROL**: This includes source modification such as acoustic treatment to machine surface, design changes, limiting operational timings, etc.
- **TRANSMISSION PATH INTERVENTION:** This includes containing the source inside a

sound insulating enclosure, constructing a noise barrier or provision of sound absorbing materials along the path.

- **RECEPTOR CONTROL:** This includes protection of the receiver by altering the work schedule or provision of personal protection devices such as ear plugs for operating noisy machinery. The measure may include dissipation and deflection methods.
- **OILING:** Proper oiling will reduce noise from the machine.

Preventive measures:

- Prescribing noise limits for vehicular traffic
- Ban on honking (usage of horns) in certain areas
- Creation of silence zones near schools and hospitals
- Redesigning buildings to make them noise proof
- Reduction of traffic density in residential areas
- Giving preference to mass public transport system.

THERMAL POLLUTION

Thermal pollution is defined as the addition of excess of undesirable heat to water thereby making it harmful to man, animal or aquatic life. Thermal pollution may also cause neither significant departures from nor activities of aquatic communities.

Sources of Thermal Pollution:

The following sources contribute to thermal pollution.

- Nuclear power plants
- Coal fired plants
- Industrial effluents
- Domestic sewage
- Hydro-electric power

Nuclear power plants: Nuclear power plants including drainage from hospitals, research institutions, nuclear experiments and explosions, discharge a lot of heat that is not utilized along with traces of toxic radio nuclides into nearby water streams. Emissions from nuclear reactors and processing installations are also responsible for increasing the temperatures of water bodies.

Coal-fired power plants: Coal fired power plants constitute a major source of thermal pollution. The condenser coils in such plants are cooled with water from nearby lakes or rivers. The resulting heated water is discharged into streams thereby raising the water temperature by 15C. The sudden fluctuation of temperature also leads to "thermal shock" killing aquatic life that has become acclimatized to living in a steady temperature.

Industrial effluents: Industries like textile, paper, pulp and sugar manufacturing release huge amounts of cooling water along with effluents into nearby natural water bodies. The waters polluted by sudden and heavy organic loads result in severe drop in levels of dissolved oxygen leading to death of several aquatic organisms.

Domestic Sewage: Domestic sewage is discharged into rivers, lakes, canals or streams with minimal treatment or without any treatment. These wastes have a higher organic temperature and organic load. This leads to decrease in dissolved oxygen content in the receiving waters resulting in the set-up of anaerobic conditions causing release of foul and offensive gases in water

Hydro-electric power: Generation of hydroelectric power sometimes leads to negative thermal loading in water systems. Apart from electric power industries, various factories with cooling requirement contribute to thermal loading.

Effects of Thermal pollution

- Reduction in dissolved oxygen: Concentration of Dissolved Oxygen (DO) decreases with increase in temperature.
- Increase in toxicity: The rising temperature increases the toxicity of the poison present in water
- Interference in biological activity: Temperature is considered to be of vital significance to physiology, metabolism and biochemical processes that control respiratory rates, digestion, excretion, and overall development of aquatic organisms. Temperature changes cause total disruption to the entire ecosystem.
- Interference in reproduction: In fishes, several activities like nest building, spawning, hatching, migration and reproduction depend on optimum temperature.
- Direct mortality: Thermal pollution is directly responsible for mortality of aquatic organisms. Increase in temperature of water leads to exhaustion of microorganisms thereby shortening the life span of fish
- Food storage for fish: Abrupt changes in temperature alter the seasonal variation in the type and abundance of lower organisms leading to shortage of right food for fish at the right time.

Control Measures For Thermal Pollution

The following methods can be adapted to control high temperature caused by thermal discharges:

Cooling towers: Use of water from water systems for cooling systems for cooling purposes, with subsequent return to the water way after passage through a condenser, is called cooling process. Cooling towers transfer heat from hot water to the atmosphere by evaporation. Cooling towers are of two types:

Wet cooling tower: Hot water coming out from the condenser (reactor) is allowed to spray over baffles. Cool air, with high velocity, is passed from sides, which takes away the heat and cools the water.

Dry cooling tower: Here, hot water is allowed to flow in long spiral pipes. Cool air with the help of a fan is passed over these hot pipes, which cools down hot water. This cool water can be recycled.

Cooling ponds: Cooling ponds are the best way to cool thermal discharges. Heated effluents on the surface of the water in cooling ponds maximize dissipation of heat to the atmosphere and minimize the water area and volume. The warm water wedge acts like a cooling pond.

Spray ponds: The water coming out from condensers is allowed to pass into the ponds through sprayers. Here water is sprayed through nozzles as fine droplets. Heat from the fine droplets gets dissipated to the atmosphere.

Artificial lakes: Artificial lakes are man-made water bodies that offer once-through cooling. The heated effluents can be discharged into the lake at one end and water for cooling purposes may be withdrawn from the other end. The heat is eventually dissipated through evaporation

Nuclear Hazards

The radiation hazard in the environment comes from ultraviolet, visible, cosmic rays and micro wave radiation which produces genetic mutation in man.

NUCLEAR HAZARDS

The radiation hazard in the environment comes from ultraviolet, visible, cosmic rays and micro wave radiation which produces genetic mutation in man.

Sources of Nuclear Hazards

Natural Sources – This is in space which emits cosmic rays.

Man-made Sources – (Anthropogenic sources) these are nuclear power plants, X-rays, nuclear accidents, nuclear bombs, diagnostic kits.

Effects of Nuclear Hazards

- Exposure of the brain and central nervous system to high doses of radiation causes delirium, convulsions and death within hours or days.
- The use of eye is vulnerable to radiation. As its cell die, they become opaque forming cataracts that impair sight.
- Acute radiation sickness is marked by vomiting; bleeding of gums and in severe cases mouth ulcers.
- Nausea and vomiting often begin a few hours after the gastrointestinal tract is exposed. Infection of the intestinal wall can kill weeks afterwards.
- Unborn children are vulnerable to brain damage or mental retardation, especially if
- Irradiation occurs during formation of the central nervous system in early pregnancy.

Control measures

- Nuclear devices should never be exploded in air.
- In nuclear reactors, closed cycle coolant system with gaseous coolant may be used to prevent extraneous activation products.
- Containments may also be employed to decrease the radioactive emissions.
- Extreme care should be exercised in the disposal of industrial wastes contaminated with radio nuclides.
- Use of high chimneys and ventilations at the working place where radioactive contamination is high. It seems to be an effective way for dispersing pollutants.

SOLID WASTE MANAGEMENT: CAUSES, EFFECTS AND CONTROL MEASURES OF URBAN AND INDUSTRIAL WASTES.

SOLID WASTE MANAGEMENT

- Rapid population growth and urbanization in developing countries has led to people generating enormous quantities of solid waste and consequent environmental degradation.
- The waste is normally disposed in open dumps creating nuisance and environmental degradation.
- Solid wastes cause a major risk to public health and the environment. Management of solid

INCINERATION:

In this method municipal solid wastes are burnt in a furnace called incinerator. Combustible substances such as rubbish, garbage, dead organisms and non-combustible matter such as glass, porcelain and metals are separated before feeding to incinerators. The non-combustible materials can be left out for recycling and reuse. The leftover ashes and clinkers may account for about 10 to 20% which need further disposal by sanitary landfill or some other means.

ADVANTAGES

- Residue is only 20-25% of the original and can be used as clinker after treatment
- Requires very little space

- Cost of transportation is not high if the incinerator is located within city limits
- Safest from hygienic point of view
- An incinerator plant of 3000 tons per day capacity can generate 3MW of power.

COMPOSTING

In this method, bulk organic waste is converted into fertilizer by biological action.

Separated compostable waste is dumped in underground trenches in layers of 1.5m and finally covered with earth of 20cm and left for decomposition. Sometimes, actinomycetes are introduced for active decomposition.

Advantages

- Manure added to soil increases water retention and ion-exchange capacity of soil.
- This method can be used to treat several industrial solid wastes.
- Manure can be sold thereby reducing cost of disposing wastes
- Recycling can be done
- wastes is important in order to minimize the adverse effects posed by their indiscriminate disposal.

TYPES OF SOLID WASTES:

Depending on the nature of origin, solid wastes are classified into

- Urban or municipal wastes
- Industrial wastes and
- Hazardous wastes
- Sources of urban wastes

Urban wastes include the following wastes:

Domestic wastes containing a variety of materials thrown out from homes

Ex: Food waste, Cloth, Waste paper, Glass bottles, Polythene bags, Waste metals, etc.

Commercial wastes: It includes wastes coming out from shops, markets, hotels, offices, institutions, etc.

Ex: Waste paper, packaging material, cans, bottle, polythene bags, etc.

Construction wastes: It includes wastes of construction materials.

Ex: Wood, Concrete, Debris, etc.

Biomedical wastes: It includes mostly waste organic materials

Ex: Anatomical wastes, Infectious wastes, etc.

Classification of urban wastes

Urban wastes are classified into:

- **Bio-degradable wastes -** Those wastes that can be degraded by micro-organisms are called bio-degradable wastes. Ex: Food, vegetables, tea leaves, dry leaves, etc.
- **Non-biodegradable wastes:** Urban solid waste materials that cannot be degraded by micro- organisms are called non-biodegradable wastes. Ex: Polythene bags, scrap materials, glass bottles, etc.

SOURCES OF INDUSTRIAL WASTES

The main source of industrial wastes is chemical industries, metal and mineral processing industries.

Ex: Nuclear plants: It generated radioactive wastes

Thermal power plants: It produces fly ash in large quantities

Chemical Industries: It produces large quantities of hazardous and toxic materials.

Other industries: Other industries produce packing materials, rubbish, organic wastes, acid, alkali, scrap metals, rubber, plastic, paper, glass, wood, oils, paints, dyes, etc.

EFFECT OF IMPROPER SOLID WASTE MANAGEMENT

- Due to improper disposal of municipal solid waste on the roads and immediate surroundings, biodegradable materials undergo decomposition producing foul smell and become a breeding ground for disease vectors.
- Industrial solid wastes are the source for toxic metals and hazardous wastes that affect soil characteristics and productivity of soils when they are dumped on the soil
- Toxic substances may percolate into the ground and contaminate the groundwater.
- Burning of industrial or domestic wastes (cans, pesticides, plastics, radioactive materials and batteries) produce furans, dioxins and polychlorinated biphenyls that are harmful to human beings.
- Solid waste management involves waste generation, mode of collection, transportation, segregation of wastes and disposal techniques.

STEPS INVOLVED IN SOLID WASTE MANAGEMENT:

Two important steps involved in solid waste management are:

A. Reduce, Reuse and Recycle of Raw Materials B. Discarding wastes

Reduce - If usage of raw materials is reduced, the generation of waste also gets reduced.

Reuse - Refillable containers that are discarded after use can be reused.

Rubber rings can be made from discarded cycle tubes and this reduces waste generation duringmanufacture of rubber bands.

Recycle- Recycling is the reprocessing of discarded materials into new useful products.

Ex: Old aluminum cans and glass bottles are melted and recast into new cans and bottles

Preparation of cellulose insulation from paper.

Preparation of automobile body and construction material from steel cans.

This method (Reduce, Reuse & Recycle), i.e, 3R's help save money, energy, raw materials and reduces pollution.

B. DISCARDING WASTES:

The following methods are adopted for discarding wastes:

A. Landfill
B. Incineration and
C. Composting

A. LAND FILL: Solid wastes are placed in a sanitary landfill in which alternate layers of 80 cm thick refuse is covered with selected earth-fill of 20 cm thickness. After 2-3 years solid waste volume shrinks by 25-30% and land is used for parks, roads and small buildings. This is the most common and cheapest method of waste disposal and is mostly employed in Indian cities.

Advantages:

- It is simple and economical
- Segregation of wastes is not required
- Land filled areas can be reclaimed and used for other purposes
- Converts low-lying, marshy waste-land into useful areas.
- Natural resources are returned to soil and recycled.

ROLE OF AN INDIVIDUAL IN PREVENTION OF POLLUTION

- Individuals should minimize wastage of resources such as electricity. Every unit of electricity saved is equivalent unit of electricity produced as it not only saves the fuel that would be used to produce that electricity, but also help to prevent pollution that is accompanied by burning of that fuel. Therefore, person should always switch off appliances when not in use.

- Individuals should prefer walking or use cycles instead of using motor vehicles, especially when distances to be travelled are small.
- Individuals can make considerable contribution by using mass transport (buses, trains, etc) instead of using personal vehicles.
- When going to workplace, colleagues from nearby localities should pool vehicles instead of going in individual personal vehicles.
- Taking personal vehicles for periodic pollution checks at centers approved by authorities.
- Individuals should reuse items whenever possible.
- Products that are made of recycled material should be given preference.
- Use gunny bags made of jute instead of plastic bags.
- Take part in environment conservation drives such as tree planting drives.
- Use water resources efficiently.
- Use renewable resources by installing equipment such as solar heaters and using solar cookers.

DISASTER MANAGEMENT: FLOODS, EARTH QUAKE, CYCLONE AND LANDSLIDES.

FLOODS:- Increased rainfall or rapid snow melting causes more flow of water in the streams. This excess water flow in a stream covering the adjacent land is called a flood. Floodplain is defined in terms of a flood frequency. Flood frequency is referred as 10 -year flood, 100- year flood, etc. A 10-year flood at any point in a stream is that discharge of water which may be expected to occur on average once in 10 years. Floodplains are generally fertile, flat and easily formed.

CAUSES OF FLOOD

- Construction of buildings in a floodplain
- Removing vegetation
- Paving roads and parking areas
- Deforestation
- Heavy rainfall
- Urbanization
- Earthquake.

Effects of Flood

- Erosion of top soil and vegetation
- Damage and loss to land, house and property
- Spread of endemic water borne diseases
- Interruption of basic facilities of community such as highways, railways, telephone, electricity and day-to-day essentials
- Silting of reservoirs and dams

FLOOD CONTROL

- Construction of flood control dam
- Deepening, widening and straightening of streams
- Lining of streams
- Banning of construction of buildings in flood plains
- Converting flood-plains into wildlife habitat, parks, and recreation areas.

LANDSLIDES

- Landslides occur when mass of earth material move downward. It is also called mass wasting or mass movement.
- Sudden landslide occurs when unconsolidated sediments of a hillside are saturated by rainfall or water logging.
- Many landslides take place in coincidence with earthquakes. The most common form of landslides is earthquake induced landslides or more specifically rock falls and slides of rock fragments that form on steeps lopes.
- The size of area affected by earthquake induced landslides depends on the magnitude of the earthquake, its focal depth, the topography and geologic conditions near the causative fault, the amplitude, frequency, composition and duration of ground shaking.

Control measures for land slides

- Avoid construction activity in landslide occurring areas.
- Reducing slope of hilly side
- Stabilizing the slope portion
- Increasing plantation of deep rooted vegetation on the slope.

EARTH QUAKES

- An earthquake occurs when rocks break and slip along a fault in the earth. Earthquakes occur due to deformation of crust and upper mantle of the earth.
- Due to heating and cooling of the rock below these plates, movement of adjacently overlying plates and great stresses, deformation occurs.
- Tremendous energy cans build-up between neighboring plates.
- If accumulated stress exceeds the strength of the rocks, the rocks break suddenly releasing the stored energy as an earthquake.
- The earthquake releases energy in the form of waves that radiate from the epicenter in all directions.
- The 'p' wave or primary wave alternately compresses and expands material in the same direction it is travelling.
- This wave can move through solid rocks and fluids.
- These are the fastest waves. The is wave or secondary wave is slower and shake the ground up, down, back and forth perpendicular to the direction in which it is travelling. Surface waves follow both the 'P' and 'S' waves.
- The magnitude of an earthquake is measured in Richter scale. The Richter scale is logarithmic.

Effects of earth quake

- Ground shaking
- Liquefaction of ground
- Ground displacement
- Landslides
- Flood
- Fire

Control of earthquake

- There is virtually no technique to control the occurrence of earthquake. However, certain preventive measures can be taken to minimize the damage.
- Minimizing development activity (especially construction, mining, construction of dams and reservoirs) in areas known to be active seismic zones.
- Continuously monitoring seismic activity using 'seismographs' and alerting people regarding any recorded disturbance in advance.

Expected Questions

1. Short Answer Type Question:

a) What do you mean by environmental pollution?
b) Name different types of environmental pollution.
c) What are pollutants? Mention different types of pollutants.
d) What are air pollutants?
e) What are the causes of air pollution?
f) What is the effect of air pollution on human health?
g) How is air pollution controlled?
h) Mention the causes of thermal pollution.
i) Mention the causes of nuclear pollution.
j) What are the water pollutants and classify them?
k) Discuss different categories of water pollution.
l) What are the causes of water pollution?
m) What is soil pollution? Discuss the types of soil pollution.
n) Discuss the causes of soil pollution.
o) What is meant by noise pollution? Give the source/causes of noise pollution.
p) Discuss the effect of noise pollution.
q) What is noise? What are the acceptable noise levels for various occupations?
r) What is a solid waste? What is meant by solid waste management?
s) What is disaster management? Mention the types of environmental disaster.
t) Give the causes of cyclone. What are the effects of cyclone?

2. Long Answer type Question:

a) What is air pollution? Describe the causes, consequences and control measures.
b) What is noise pollution? Describe its causes, effects, and prevention, and control measures.
c) Describe the effects of soil pollution and its control.
d) What is water pollution? Describe its causes, effects, and prevention, and control measures.
e) Discuss the types of environmental pollution.
f) Discuss the various sources of marine pollution. What are the effects and control measures of marine pollution?
g) Write an essay on floods, their control and management.
h) Give a detailed the four aspects of management of environmental disasters.
i) What are the measures you would take in case of floods and cyclones?
j) Write short notes on nuclear hazards.
k) What are the various methods of solid waste treatment? Describe any one method.

GANDHI ACADEMY OF TECHNOLOGY AND ENGINEERING

GOLANTHARA, BERHAMPUR

LECTURE NOTES

ON

ENERGY CONVERSION - I

For 4th Semester

Electrical Engineering

(As per Syllabus prescribed by SCTE&VT, Odisha)

Prepared By

Mrs. Soumya Shyamali Mahapatra

(Lecturer in Electrical Engineering)

Th1. ENERGY CONVERSION – I

Name of the Course: Diploma in Electrical Engineering			
Course code:		Semester	4th
Total Period:	75 (60L + 15T)	Examination	3 hrs
Theory periods:	4P / week	Internal Assessment :	20
Tutorial:	1 P / week		
Maximum marks:	100	End Semester examination:	80

A. RATIONALE

Energy Conversion-I deals with DC machines and transformers. The application of DC generators and motors in modern industries are still in practice. The electrical technicians have to look after the installation, operation, maintenance and control of such machine. So the knowledge of these machines is felt essential. Transformers of various voltage ratios and KVA ratings are in wide use in industries as well as in distribution and transmission.

B. OBJECTIVES

After completion of this subject the student will be able to:

1. To acquire knowledge of construction, characteristic and control of the DC machines.
2. To acquire knowledge on performance of DC machines and transformers.
3. To acquire knowledge of testing and maintenance of transformers and DC machines.

C. TOPIC WISE DISTRIBUTION OF PERIODS		
Sl. No.	**Topic**	**Periods**
1.	DC GENERATORS	17
2.	DC MOTORS	15
3.	SINGLE PHASE TRANSFORMER	20
4.	AUTO TRANSFORMER	03
5.	INSTRUMENT TRANSFORMERS	05
TOTAL		60

D. COURSE CONTENT IN TERMS OF SPECIFIC OBJECTIVES

1. D.C GENERATOR

1.1. Operating principle of generator
1.2. Constructional features of DC machine.
  1.2.1. Yoke, Pole & field winding, Armature, Commutator.
  1.2.2. Armature winding, back pitch, Front pitch, Resultant pitch and commutator- pitch.
  1.2.3. Simple Lap and wave winding, Dummy coils.
1.3. Different types of D.C. machines (Shunt, Series and Compound)
1.4. Derivation of EMF equation of DC generators. (Solve problems)
1.5. Losses and efficiency of DC generator. Condition for maximum efficiency and numerical problems.

1.6. Armature reaction in D.C. machine
1.7. Commutation and methods of improving commutation.
 1.7.1. Role of inter poles and compensating winding in commutation.
1.8. Characteristics of D.C. Generators
1.9. Application of different types of D.C. Generators.
1.10. Concept of critical resistance and critical speed of DC shunt generator
1.11. Conditions of Build-up of emf of DC generator.
1.12. Parallel operation of D.C. Generators.
1.13. Uses of D.C generators.

2. D. C. MOTORS

2.1. Basic working principle of DC motor
2.2. Significance of back emf in D.C. Motor.
2.3. Voltage equation of D.C. Motor and condition for maximum power output(simple problems)
2.4. Derive torque equation (solve problems)
2.5. Characteristics of shunt, series and compound motors and their application.
2.6. Starting method of shunt, series and compound motors.
2.7. Speed control of D.C shunt motors by Flux control method. Armature voltage Control method. Solve problems
2.8. Speed control of D.C. series motors by Field Flux control method, Tapped field method and series-parallel method
2.9. Determination of efficiency of D.C. Machine by Brake test method(solve numerical problems)
2.10. Determination of efficiency of D.C. Machine by Swinburne's Test method(solve numerical problems)
2.11. Losses, efficiency and power stages of D.C. motor(solve numerical problems)
2.12. Uses of D.C. motors

3. SINGLE PHASE TRANSFORMER

3.1 Working principle of transformer.
3.2 Constructional feature of Transformer.
 3.2.1 Arrangement of core & winding in different types of transformer.
 3.2.2 Brief ideas about transformer accessories such as conservator, tank, breather, and explosion vent etc.
 3.2.3 Explain types of cooling methods
3.3 State the procedures for Care and maintenance.
3.4 EMF equation of transformer.
3.5 Ideal transformer voltage transformation ratio
3.6 Operation of Transformer at no load, on load with phasor diagrams.
3.7 Equivalent Resistance, Leakage Reactance and Impedance of transformer.
3.8 To draw phasor diagram of transformer on load, with winding Resistance and Magnetic leakage with using upf, leading pf and lagging pf load.
3.9 To explain Equivalent circuit and solve numerical problems.
3.10 Approximate & exact voltage drop calculation of a Transformer.
3.11 Regulation of transformer.
3.12 Different types of losses in a Transformer. Explain Open circuit and Short Circuit test.(Solve numerical problems)
3.13 Explain Efficiency, efficiency at different loads and power factors, condition for maximum efficiency (solve problems)
3.14 Explain All Day Efficiency (solve problems)
3.15 Determination of load corresponding to Maximum efficiency.
3.16 Parallel operation of single phase transformer.

4. AUTO TRANSFORMER

4.1. Constructional features of Auto transformer.
4.2. Working principle of single phase Auto Transformer.
4.3. Comparison of Auto transformer with an two winding transformer (saving of Copper).
4.4. Uses of Auto transformer.
4.5. Explain Tap changer with transformer (on load and off load condition)

5. INSTRUMENT TRANSFORMERS

1.1 Explain Current Transformer and Potential Transformer

1.2 Define Ratio error, Phase angle error, Burden.

1.3 Uses of C.T. and P.T.

<u>Syllabus coverage up to Internal assessment</u>

Chapters: 1 and 2.

Learning Resources:

Sl.No	***Title of the Book***	***Name of Author***	***Publisher***
1	*Electrical Technology – II*	*B. L. Thareja and A. K. Thareja*	*S.Chand*
2	*A Textbook of Electrical Machines*	*K R Siddhapura, D B Raval*	*Vikas*
3.	*Electrical Technology*	*J. B. Gupta*	*S.K.Kataria and Sons*
4.	*Electric Machine*	*Ashfaq Husain*	*Dhanpat Rai and Sons*
5.	*Electrical Machine*	*S. K. Bhattarcharya*	*TMH*
6.	*Electrical Machines*	*D P Kothari, I J Nagrath*	*Mc Graw Hill*
7	*Electrical Machines*	*Prithwiraj purakait and Indrayudh Bandyopadhyay*	*OXFORD*

DC Generators

Principle of operation of DC Generator

A D.C generator as shown in figure below the armature be driven by a prime mover in the clock wise direction and the stator field is excited to produce the field poles as shown. There will be induced voltage in each armature conductor. The direction of the induced voltage can be determined by applying *Fleming's right hand rule.* All the conductors under the influence of North Pole will have $\otimes$ directed induced voltage, while the conductors under the influence of South Pole will have $\odot$ induced voltage in them. For a loaded generator the direction of the armature current will be same as that of the induced voltages. Thus $\otimes$ and $\odot$ also represent the direction of the currents in the conductors. We know, a current carrying conductor placed in a magnetic field experiences force, the direction of which can be obtained by applying *Fleming's left hand rule.* Applying this rule to the armature conductors in fig 1.9, the rotor experiences a torque (T_e) in the counter clockwise direction (i.e., opposite to the direction of rotation) known as back torque. For steady speed operation back torque is equal to the machines input torque (T_{pm}) i.e. the torque supplied by prime mover.

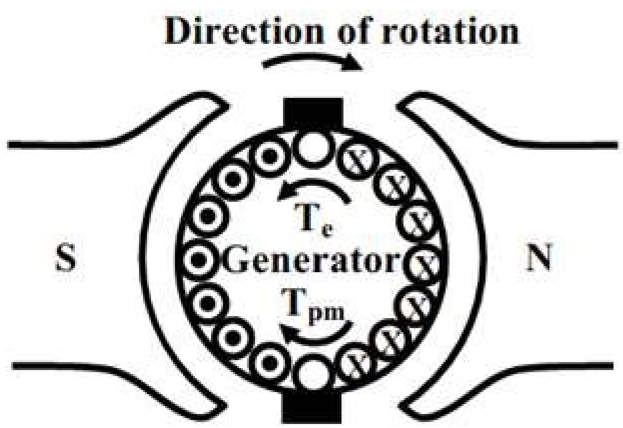


Fig. 1.9 Action of DC generator

Action of Commutator

In DC machines the current in each wire of the armature is actually alternating, and hence a device is required to convert the alternating current generated in the DC generator by electromagnetic induction into direct current, or at the armature of a DC motor to convert the input direct current into alternating current at appropriate times, as illustrated in Fig. 1.10.

DC generator: induced AC *emf* is converted to DC voltage;

DC motor: input direct current is converted to alternating current in the armature at appropriate times to produce a unidirectional torque. The commutator consists of insulated copper segments mounted on an insulated tube. Armature coils are connected in series through the commutator segments. Two brushes are pressed to the commutator to permit current flow. The brushes are placed in the neutral zone, where the magnetic field is close to zero, to reduce arcing. The *commutator* switches the current from one rotor coil to the adjacent coil. The switching requires the interruption of the coil current. The sudden interruption of an inductive current generates high voltages. The high voltage produces flashover and arcing between the commutator segment and the brush.

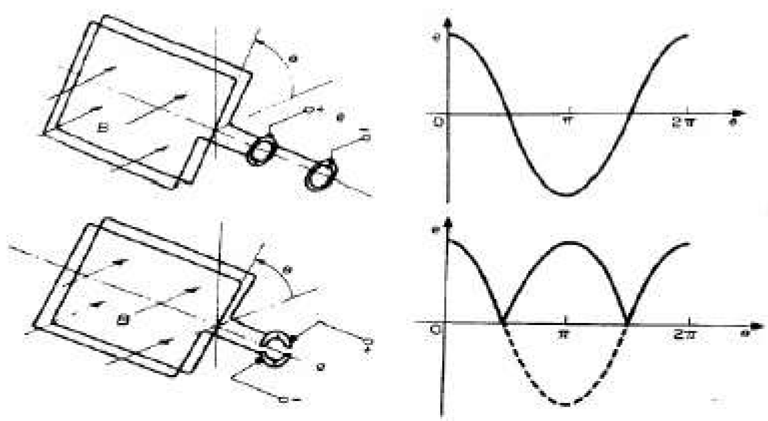


Fig. 1.10Action of Commutator

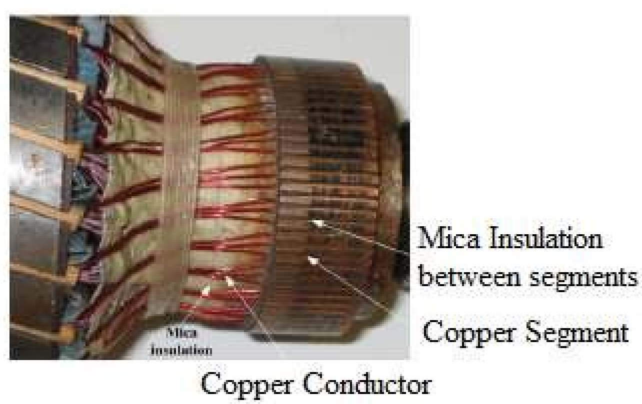


Fig. 1.11 Mechanical view of commutator

Constructional Features

The stator of the dc machine has poles, which are excited by dc current to produce magnetic fields. In the neutral zone, in the middle between the poles, commutating poles or interpoles are placed to reduce sparking of the commutator due to armature reaction. The commutating poles are supplied by dc current. Compensating windings are mounted on the main poles. Field poles are mounted on an iron core that provides a closed magnetic circuit. The motor housing supports the iron core, the brushes and the bearings. The rotor has a ring-shaped laminated iron core with slots. Coils with several turns are placed in the slots. The distance between the two legs of the coil is about 180 electric degrees for full pitch. The coils are connected in series through the commutator segments. The ends of each coil are connected to a commutator segment.

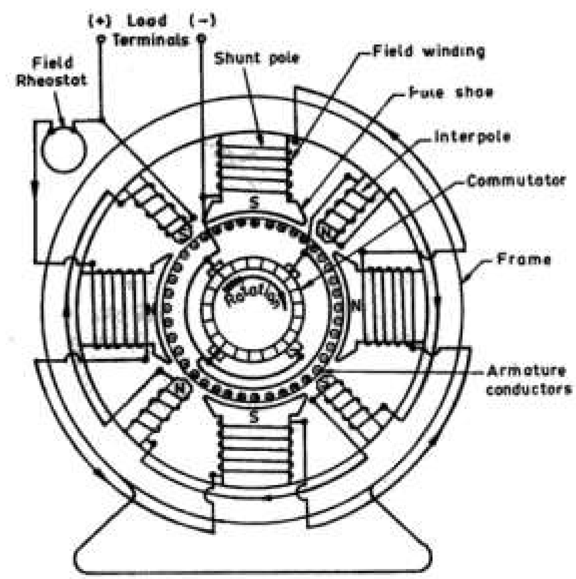


Fig. 1.12 Cross sectional view of DC Machine

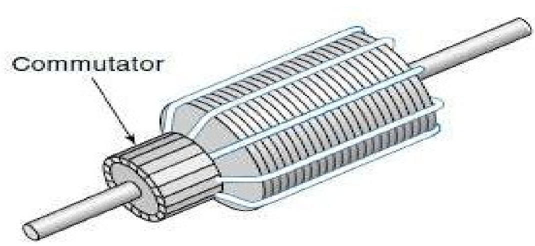


Fig. 1.13 Armature of DC Machine

Armature Winding

DC machines armature consists of armature conductors. The conductors distributed in slots provided on the periphery of the armature is called armature winding. Depending on the way in which the coils are interconnected at the commutator end of the armature, the windings can be classified as lap and wave windings. Further they can be classified as simplex and multiplex.

Coil Span/Coil Pitch:

It represents the span of the coil. For full pitched winding, the span is 180^0 electrical or number of slots per pole. Coil pitch can be represented in terms of electrical degrees, slots or conductor. A full pitched coil leads to maximum voltage per coil.

Back Pitch *(Yb)*:

It is the distance measured in between the two coil sides of the same coil at the back end of the armature, the commutator end being the front end of armature. It can be represented in terms of number of slots or coil sides. Back pitch also represents the span of coil.

Front Pitch (*Yf*):

The distance between the two coil sides of two different coils connected in series at the front end of the armature is called front pitch.

Lap Winding

Lap winding is suitable for low voltage high current machines because of more number of parallel paths. The number of parallel path in lap winding is equal to number of poles.

$$A=P$$

Equalizing rings are connected in lap winding.

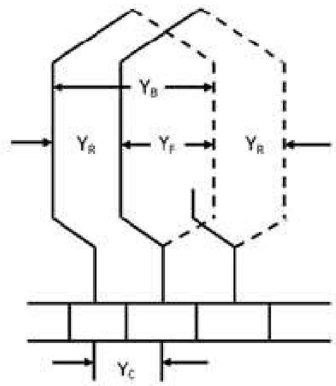


Fig. 1.14 Winding diagram of lap winding

Wave Winding

Wave winding is used for high voltage low current machines. In case of wave winding, the number of parallel path (A) = 2 irrespective of number of poles. Each path will have conductors connected in series.

Equalizing rings are not required in wave winding.

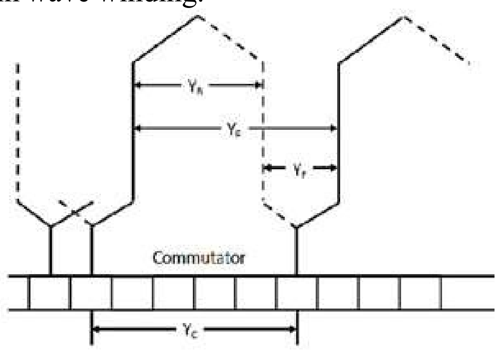


Fig. 1.15 Winding diagram of wave winding

Simplex and Multiplex Winding

Fig 1.14 and Fig. 1.15 shows simplex lap and simplex wave winding.

The degree of multiplicity of a multiplex winding indicates the relative number of parallel paths with respect to the number of parallel paths in the corresponding simplex winding. For example a duplex lap or wave winding is a lap or wave winding having twice as many as parallel paths as a simplex lap or wave winding respectively. The winding can be triplex or quadruplex winding in similar manner.

Use of Laminated Armature

The armature winding of DC machine should be laminated to reduce eddy current losses. The armature body (rotor) rotates in the field magnetic field. Thus in the core of the armature voltage induced which in turn causes current to flow in the body. This current is known as eddy current. This current causes loss and thus heat will be generated. This loss depends on the amount of current flow. To reduce the amount of current flow the resistance of the body should be increased. Thus using lamination the resistance of the path through which current flows will be increased. The amount of eddy current will be reduced and thus eddy current loss can be minimized.

EMF Equation

Let ϕ = flux per pole in weber

Z = number of armature conductors = Number of slots X conductors per slot.

P = Number of poles; A= Number of parallel paths in armature.

A= P for lap wound armature; A=2 for wave wound armature

N = speed of armature in rpm; E = induced emf in each parallel path.

Average emf generated/conductor in one revolution = $\frac{d\phi}{dt}$

Flux cut by a conductor in one revolution = $d\phi = P\phi$ weber.

Since Number of revolutions/second= $\frac{N}{60}$

Time taken for one revolution = $dt = \frac{60}{N}$ seconds

EMF generated/conductor = $\frac{d\phi}{dt} = \frac{P\phi}{\frac{60}{N}} = \frac{P\phi N}{60}$

Since each path has $\frac{Z}{A}$ conductors in series,

EMF generated in each path is $E = \frac{P\phi N}{60} \times \frac{Z}{A}$

$$\boxed{E = \frac{P\phi ZN}{60A}}$$

Methods of excitation

DC machines are excited in two ways-

Separate excitation:

When the field winding is connected to an external source to produce field flux. According to the type of excitation this machines are called separately excited dc machine.

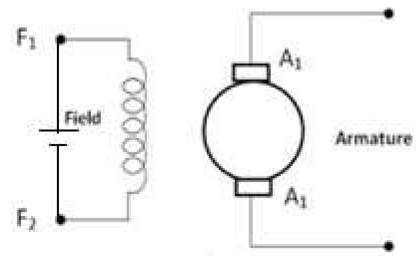


Fig. 1.16 Schematic diagram of separately excited dc machine

Self-excitation:

When the field winding is connected with the armature to produce field flux. A self-excited machine requires residual magnetism for operation. According to the type of excitation this machines are called self-excited dc machine.

Depending on the type of field winding connection DC machines can be classified as:

Shunt machine:

The field winding consisting of large number of turns of thin wire is usually excited in parallel with armature circuit and hence the name shunt field winding. This winding will be having more resistance and hence carries less current.

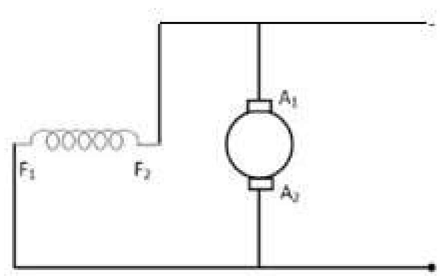


Fig. 1.17 Schematic diagram of dc shunt machine

Series machine:

The field winding has a few turns of thick wire and is connected in series with armature.

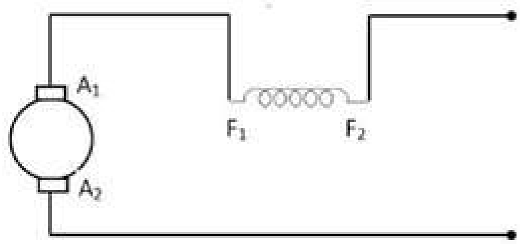
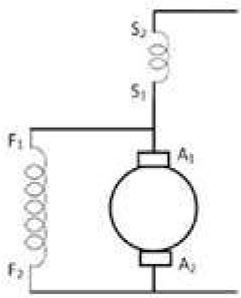


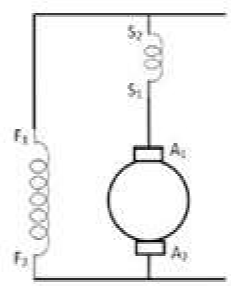
Fig.1.18 Schematic diagram of dc series machine

Compound machine:

Compound wound machine comprises of both series and shunt windings and can be either short shunt or long shunt, cumulative, differential or flat compounded.



A



B

Fig. 1.19 A A Schematic diagram of short -shunt compound machine

Fig.1.19 B Schematic diagram of long-shunt compound machine

Build-up of E.M.F

When the armature is rotating with armature open circuited, an emf is induced in the armature because of the residual flux. When the field winding is connected with the armature, a current flows through the field winding (in case of shunt field winding, field current flows even on No-load and in case of series field winding only with load) and produces additional flux. This additional flux along with the residual flux generates higher voltage. This higher voltage circulates more current to generate further higher voltage. This is a cumulative process till the saturation is attained.

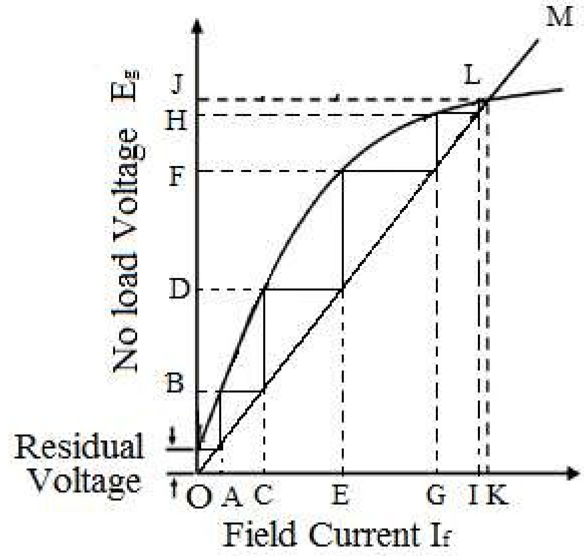


Fig. 1.20 Process of voltage build-up in DC generator

Here OM is the field resistance curve in Fig. 1.20. Initially there will be residual voltage which will create OA field current. This field current will increase the existing magnetic field and the induced voltage will increase up-to OB. This OB voltage will further applied to the field winding and increase the field current to OC. This process will continue upto the point L where the emf curve intersect with field resistance and finally the induced voltage will be OJ. This way voltage builds-up in dc generator.

1.19 Critical Resistance:-

The voltage to which it builds is decided by the resistance of the field winding as shown in the figure 1.21. If field circuit resistance is increased such that the resistance line does not cut OCC like 'OP' in the figure1.21, then the machine will fail to build up voltage to the rated value. The slope of the air gap line drawn as a tangent (OQ) to the initial linear portion of the curve represents the maximum resistance that the field circuit can have beyond which the machine fails to build up voltage. This value of field circuit resistance is called critical field resistance. The field circuit is generally designed to have a resistance value less than this so that the machine builds up the voltage to the rated value.

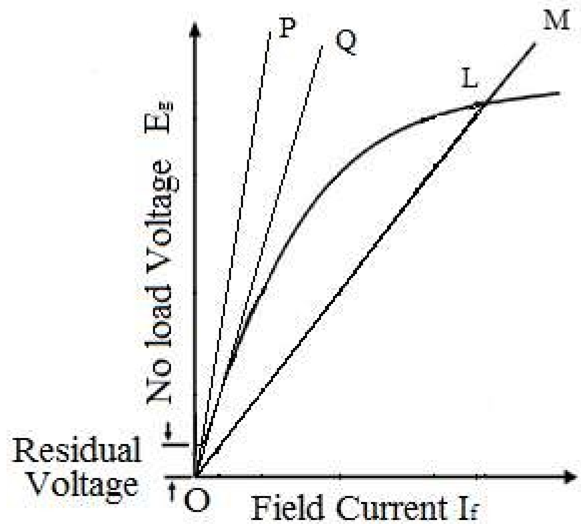
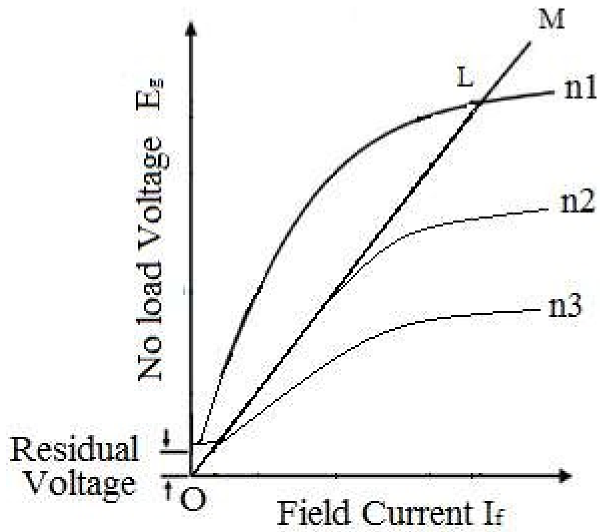


Fig. 1.21 Field current vs No-load voltage for different field resistances

Critical field resistance is defined as the maximum field circuit resistance for a given speed with which the shunt generator would excite. The shunt generator will build up voltage only if field circuit resistance is less than critical field resistance.

Critical Speed:

Voltage of a dc generator is proportional to its speed. Thus when speed will be reduced then the induced voltage will reduced. There can be such situation occur when the speed will be so low that the existing field winding resistance voltage bulid up will not occur. The speed of the generator can be lowered upto a certain level. This minimum value of the speed of the generator for which the generator can excite is called critical speed. It can also define as that speed of a generator for which the existing field resistance of generator becomes its critical field resistance.



connected to the
dual flux.

Fig. 1.22 Field current vs No-load voltage for different speed

In the above figure it is showing that when speed of the generator changes from n1 to n2 and then n3 emf production changes accordingly. Here n1>n2>n3. For speed n3 voltage build-up is not possible. The speed n2 is the critical speed. As shown in the figure at speed n2 generator field resistance become its critical field resistance.

Causes for failure to self-excite and its remedial

i. The field poles may not have residual magnetism. Then the generator will fail to excite.

   Then to restore residual magnetism field winding should be connected to an external dc voltage source. This is called flashing of field.

ii. When the direction of rotation is not proper such that flux produced by the field current reinforces the residual magnetism.

   The rotation of the machine has to be reversed.

iii. The field winding resistance is more than critical resistance then the machine will fail to excite.

   The field winding resistance should be less than critical field resistance.

iv. When the speed of the machine is less than critical speed.

   The machine's speed should be more than critical speed.

v. If the field winding connections are such that newly generated field flux is working in opposite to the existing residual magnetism. Then the generator will fail to excite.

Then the field winding connection should be reversed.

Armature Reaction

The action of magnetic field set up by armature current on the distribution of flux under main poles of a DC machine is called armature reaction.

When the armature of a DC machines carries current, the distributed armature winding produces its own mmf. The machine air gap is now acted upon by the resultant mmf distribution caused by the interaction of field ampere turns (AT_f) and armature ampere turns (AT_a). As a result the air gap flux density gets distorted.

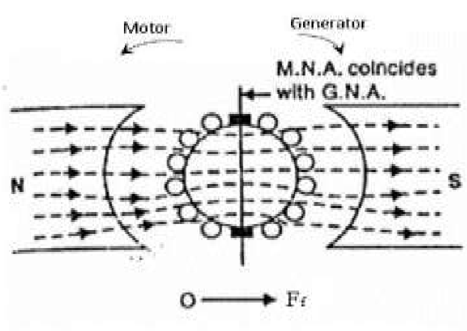


Fig. a

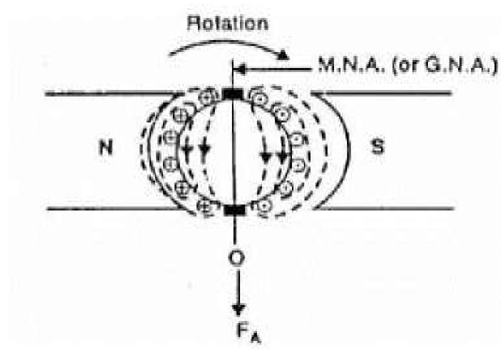


Fig. b

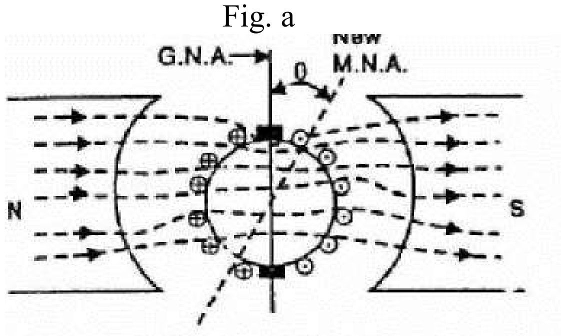


Fig. c

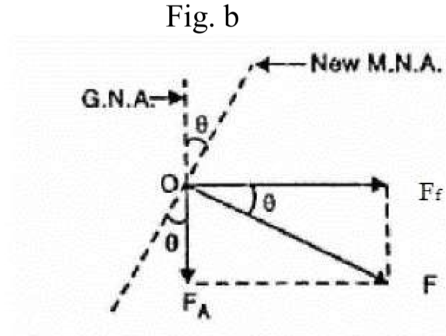


Fig. d

Figure a shows a two pole machine with single equivalent conductor in each slot and the main field mmf (F_f) acting alone. The axis of the main poles is called the direct axis (d-axis) and the interpolar axis is called quadrature axis (q-axis). It can be seen from the Figure b that armature mmf (F_a) is along the interpolar axis. F_a which is at 90^0 to the main field axis is known as cross magnetizing mmf.

Figure c shows the practical condition in which a DC machine operates when both the Field flux and armature flux are existing. Because of both fluxes are acting simultaneously, there is a shift in brush axis and crowding of flux lines at the trailing pole tip and flux lines are weakened or thinned at the leading pole tip. (The pole tip which is first met in the direction of rotation by the armature conductor is leading pole tip and the other is trailing pole tip).

If the iron in the magnetic circuit is assumed unsaturated, the net flux/pole remains unaffected by the armature reaction though the air gap flux density distribution gets distorted. If the main pole excitation is such that the iron is in the saturated region of magnetization (practical case) the increase in flux density at one end of the poles caused by armature reaction is less than the decrease at the other end, so that there is a net reduction in the flux/pole. This is called the demagnetizing effect. Thus it can be summarized that the nature of armature reaction in a DC machine is

1. Cross magnetizing with its axis along the q-axis.
2. It causes no change in flux/pole if the iron is unsaturated but causes reduction in flux/pole in the presence of iron saturation. This is termed as demagnetizing effect. The resultant mmf 'F' is shown in figure d.

1.22.1Cross Magnetizing Ampere Turns/pole(AT_c)

If the brush is shifted by an angle θ as shown in figure 1.23 then the conductors lying in between the angles BOC and DOA are carrying the current in such a way that the direction of the flux is downwards i.e., at right angles to the main flux. This results is the distortion in the main flux. Hence, these conductors are called cross magnetizing or distorting ampere conductors.

Total armature conductors/pole= $\frac{Z}{P}$

Demagnetizing conductors / pole = $Z\frac{2\theta}{360}$

Therefore cross magnetizing conductors/pole= $\frac{Z}{P} - Z\frac{2\theta}{360}$

Cross magnetizing ampere turns/pole AT$_c$= $\frac{ZI_a}{a}\left(\frac{1}{2P} - \frac{\theta}{360}\right)$

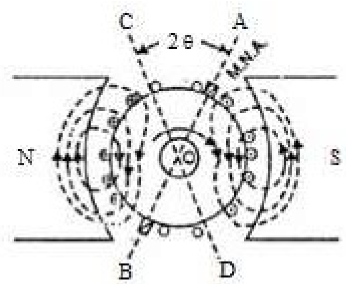


Fig. 1.23 Cross-magnetizing ampere conductors

1.22.2 Demagnetizing Ampere Turns /pole (AT$_d$)

The exact conductors which produce demagnetizing effect are shown in Fig 1.24, Where the brush axis is given a forward lead of θ so as to lie along the new axis of M.N.A. The flux produced by the current carrying conductors lying in between the angles AOC and BOD is such that, it opposes the main flux and hence they are called as demagnetizing armature conductors.

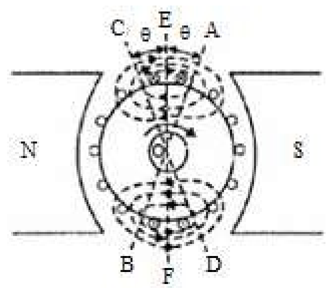


Fig. 1.24 Demagnetizing ampere conductors

Z= total no of armature conductors

Current in each armature conductors= $\frac{I_a}{a}$

θ =Forward lead in mechanical or angular deg.

Total no of armature conductors in between angles AOC & BOD =$\frac{4\theta}{360}Z$

Demagnetizing amp turns/poles AT$_d$ = $\frac{Z\theta I_a}{360a}$

Compensating Winding

Due to armature reaction flux density wave get distorted and reduced. Due to distortion of flux wave the peak flux density increases to such a high value that it creates high induced emf. If this emf is higher than the breakdown voltage across adjacent segments, a spark over could result which can easily spread over the whole commutator, and there will be a ring of fire, resulting in the complete short circuit of the armature.

To protect armature from such adverse condition armature reaction must be neutralized. To neutralize the armature reaction ampere-turns by compensating winding placed in the slots cut out in pole face such that the axis of the winding coincides with the brush axis as shown figure 1.25.

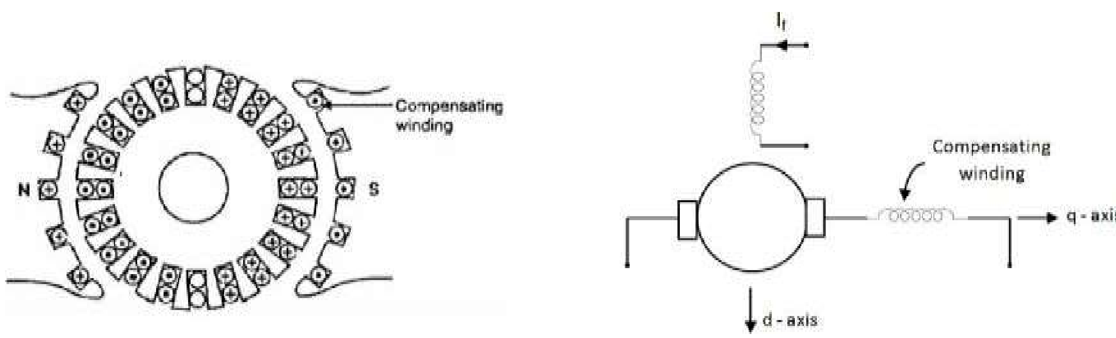


Fig. 1.25 Compensating conductors in field poles and the connection of compensating conductors with armature

The compensating windings neutralize the armature mmf directly under the pole which is the major portion because in the interpole region the air gap will be large. Compensating windings are connected in series with armature so that it will create mmf proportional to armature mmf.

The number of ampere-turns required in the compensating windings is given by

$$AT_c = \text{Total Armature Ampere Turns} \times \frac{\text{Pole Arc}}{\text{Pole Pitch}}$$

Commutation

The process of reversal of current in the short circuited armature coil is called ‘Commutation’. This process of reversal takes place when coil is passing through the interpolar axis (q-axis), the coil is short circuited through commutator segments and brush.

The process of commutation of coil ‘CD’ is shown Fig. 1.26. In sub figure ‘c’ coil ‘CD’ carries 20A current from left to right and is about to be short circuited in figure ‘d’ brush has moved by a small width and the brush current supplied by the coil are as shown. In figure ‘e’ coil ‘CD’ carries no current as the brush is at the middle of the short circuit period and the brush current in supplied by coil ‘AB’ and coil ‘EF’. In sub figure ‘f’ the coil ‘CD’ which was carrying current from left to right carries current from right to left. In sub fig ‘g’ spark is shown which is due to the reactance voltage. As the coil is embedded in the armature slots, which has high permeability, the coil possess appreciable amount of self inductance. The current is changed from +20 to –20. So due to self inductance and variation in the current from +20 to –20, a voltage is induced in the coil which is given by L dI/dt. This emf opposes the change in current in coil ‘CD’ thus sparking occurs.

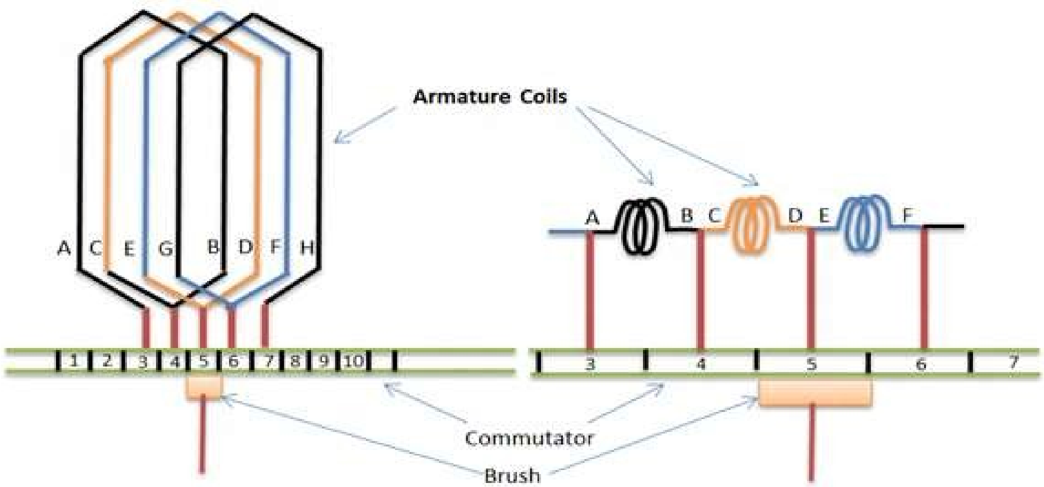


Fig a Fig b

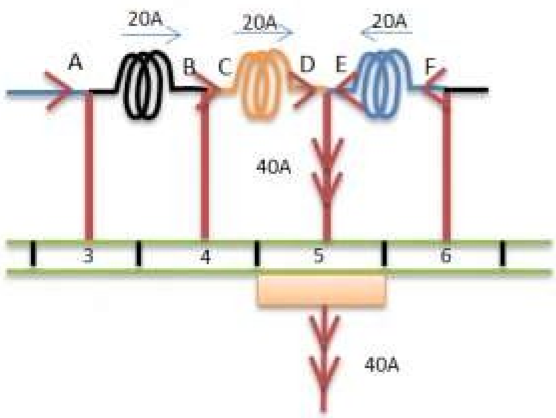


Fig.c

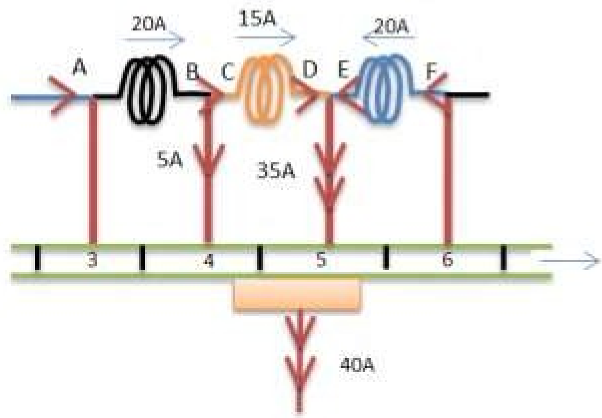


Fig.d

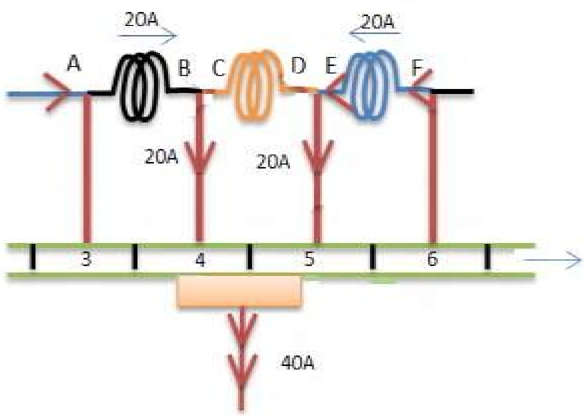


Fig.e

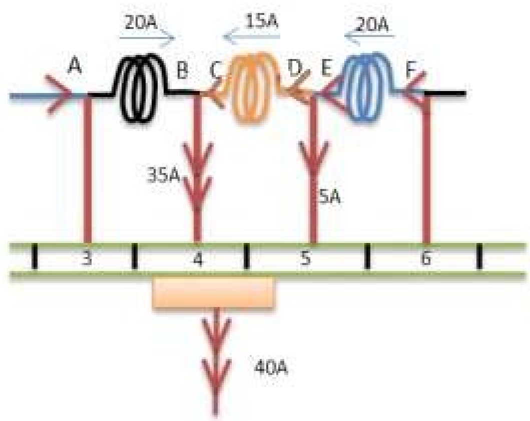


Fig.f

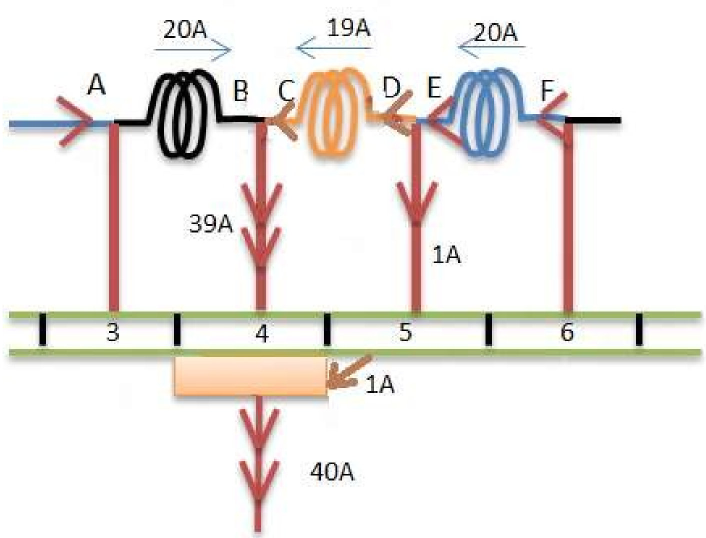


Fig.g

Fig.1.26 a-g shows the process of commutation

Reactance Voltage

During commutation sparking occurs in the commutator segment and brush due to presence of reactance voltage. This voltage is generated due to change of current in the commutating coil for its self-inductance and also due to mutual inductance of the adjacent coils. This voltage is called reactance voltage and according to Lenz's law this induced voltage oppose its cause of production. Here the cause of production is the change in current in the coil under commutation. Thus the commutation becomes poorer.

Reactance voltage = co-efficient of self-inductance X rate of change of current= $L\frac{di}{dt}$

Time of short circuit = T_c = (time required by commutator to move a distance equal to the circumferential thickness of brush)–(one mica insulating strip) = Time of commutation

Let Wb= brush width in cm

W_i = width of mica insulation in cm

V_c = peripheral velocity of commutator segments in cm/sec.

Then $T_c = \frac{W_b - W_i}{V_c}$ sec

Total change in current = I - (-I) = 2I

Therefore self-induced or reactance voltage = $L\frac{2I}{T_c}$ for linear commutation

$=1.11L\frac{2I}{T_c}$ for sinusoidal commutation

If brush width is given in terms of commutator segments, then commutator velocity should be converted in terms of commutator segments/seconds.

Method of Improving Commutation

Commutation can be improved in two ways by (i) Resistance commutation

(ii) E.M.F commutation.

Resistance Commutation

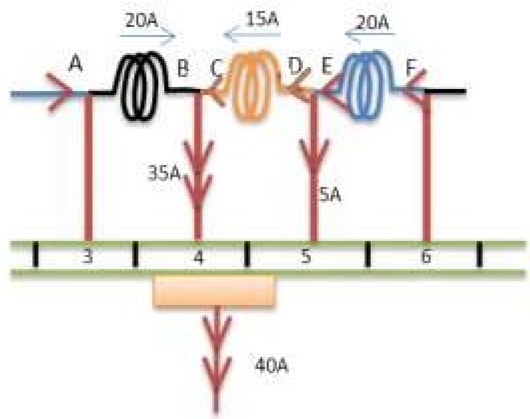


Fig. 1.27

In this method the resistance of the brushes are increased by changing then from copper brush to carbon brush. From the above figure 1.27 it is seen that when current ‘20A’ from coil ‘EF’ reaches the commutator segment ‘5’, it has two parallel paths opened to it. The first path is straight from bar ‘5’ to the brush and the other is via short circuited coil ‘CD’ to bar ‘4’ and then to brush. If copper brushes are used the current will follow the first path because of its low contact resistance. But when carbon brushes having high resistance are used, then current ‘20A’ will prefer the second path because the resistance r_1 of first path will increase due to reducing area of contact with bar ‘5’ and the resistance r_2 of second path decreases due to increasing area of contact with bar ‘4’. Hence carbon brushes help in obtaining sparkles commutation. Also, carbon brushes lubricate and polish commutator. But, because of high resistance the brush contact drop increases and the commutator has to be made larger to dissipate the heat due to loss. Carbon brushes require larger brush holders because of lower current density.

E.M.F commutation:

To neutralize sparking caused by reactance voltage in this method an emf is producied whichacts in opposite direction to that of reactance voltage, so that the reactance voltage is completely eliminated. The neutralization of emf may be done in two ways (i) by giving brush a forward lead sufficient enough to bring the short circuited coil under the influence of next pole of opposite polarity or (ii) by using interpoles or compoles. The second method is commonly employed.

Interpoles or Compoles

These are small poles fixed to the yoke and placed in between the main poles as shown in figure 1.28. They are wound with few turns of heavy gauge copper wire and are connected in series with the armature so that they carry full armature current. Their polarity in case of generator is that of the main pole ahead in the direction of rotation. Their polarity in case of motor is that of the main pole behind in the direction of rotation.

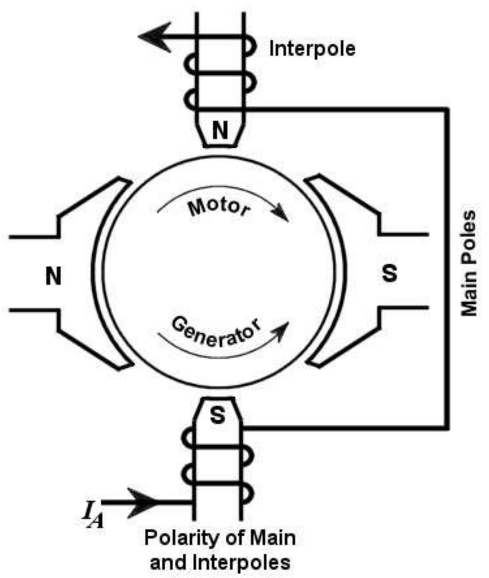


Fig. 1.28 Inter-poles of DC machines

The function of interpoles is (i) to induce an emf which is equal and opposite to that of the reactance voltage. Interpoles neutralize the cross magnetizing effect of armature reaction.

DC Machines Characteristics

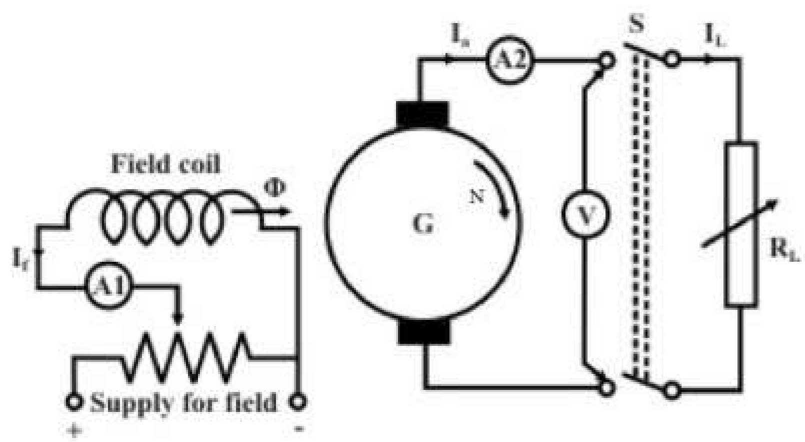
There main characteristics of dc machines are

i. No-load characteristics or Open Circuit Characteristics

ii. Load characteristics

No-load characteristics or Open Circuit Characteristics of separately excited generator

In this type of generator field winding is excited from a separate source, hence field current is independent of armature terminal voltage as shown on figure 1.29. The generator is driven by a prime mover at rated speed, say N rpm. With switch *S* in opened condition, field is excited via a potential divider connection from a separate d.c source and field current is gradually increased. The field current will establish the flux per pole ϕ. The voltmeter V connected across the armature terminals of the machine will record the generated emf ($E_g = \frac{P\phi ZN}{60A} = k\phi N$). As field current is increased, E_g will increase. E_g versus I_f plot at speed N_1, N_2, N_3 is shown in figure below. Where $N_1 > N_2 > N_3$.



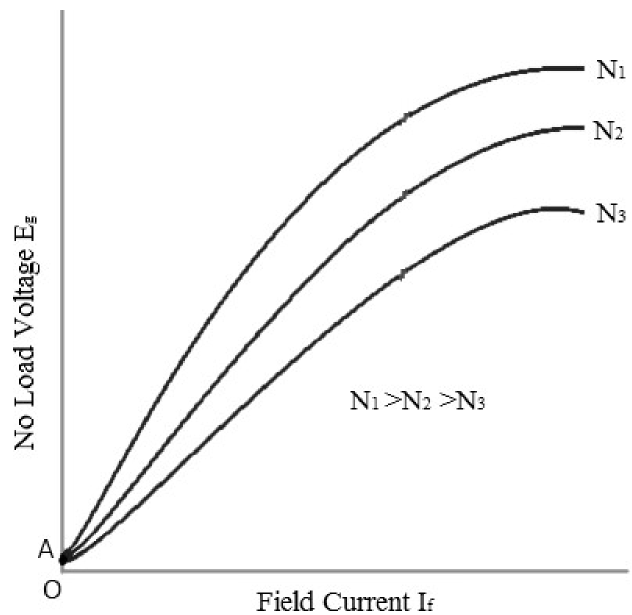


Fig. 1.29 Open circuit characteristics of DC separately excited generator

Load characteristic of separately excited generator

Load characteristic is the characteristics in between terminal voltage V_T with load current I_L of a generator with constant speed and constant field current. For $I_L = 0$, $V_T = E_g$ should be the first point on the load characteristic. With increase of load current the terminal voltage will drop due to armature resistance and reaction drop. In the figure below the rated load current shown by point A. Hence the load characteristic will be drooping in nature as shown in figure 1.30.

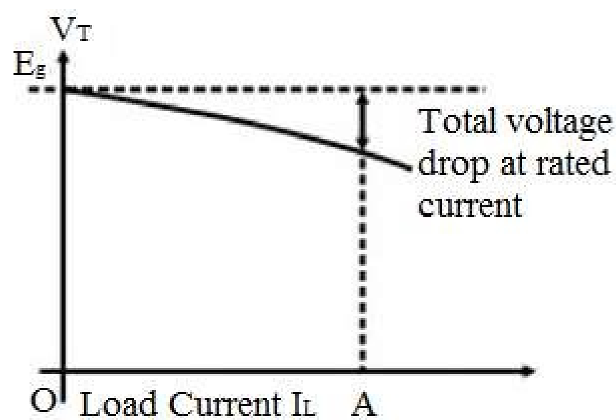


Fig 1.30 Load characteristics of DC separately excited generator

Characteristics of a shunt generator

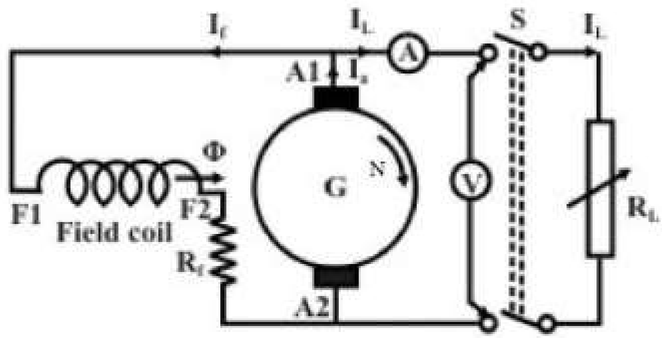


Fig 1.31 Connection diagram to obtain no-load and load characteristic

To obtain OCC of DC shunt generator the above circuit will be used with switch S kept open in fig. 1.31. As this machine is self-excited thus there is no need to use separate dc source for producing field current. The voltage build up process is described earlier. The open nature of the OCC will be similar to separately excited machine.

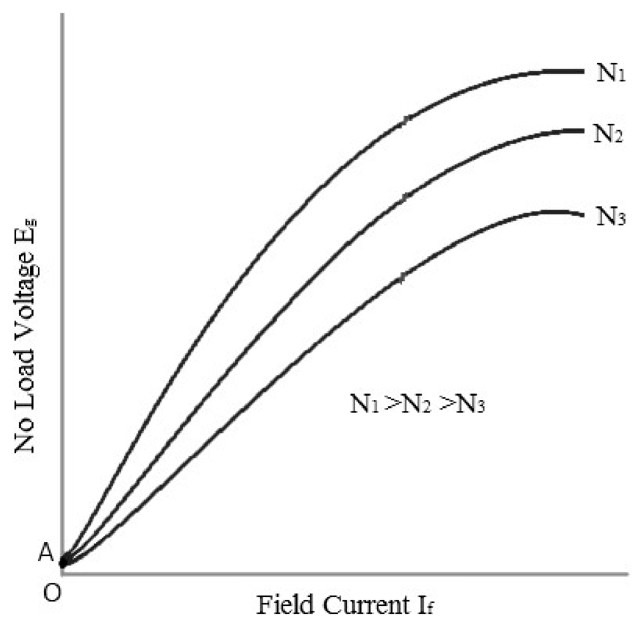


Fig. 1.32 Open circuit characteristics of DC shunt generator

Load characteristic of shunt generator

With switch S in open condition in fig. 1.31, the generator is practically under no load condition as field current is pretty small. The voltmeter reading will be E_g as shown in figures below. In other words, $V_T = E_g$, $I_L = 0$ is the first point in the load characteristic. To load the machine S is closed and the load resistances decreased so that it delivers load current I_L. Unlike separately excited motor, here $I_L \neq I_a$. In fact, for shunt generator, $I_a = I_L - I_f$. So increase of I_L will mean increase of I_a as well. The drop in the terminal voltage will be caused by the usual $I_a r_a$ drop, brush voltage drop and armature reaction effect. Apart from these, in shunt generator, as terminal voltage decreases, field current hence ϕ also decreases causing additional drop in terminal voltage. Remember in shunt generator, field current is decided by the terminal voltage by virtue of its parallel connection with the armature. Figure 1.33 gives the plot of terminal voltage versus load current which is called the load characteristic.

As the load resistance is decreased (load current increased), the terminal voltage drops until point B is reached. If load resistance is further decreased, the load current increases momentarily. This momentary increase in load current produces more armature reaction thus causing a reduction in the terminal voltage and field current. The net reduction in terminal voltage is so large that the load current decreases and the characteristic turns back. In case the machine is short circuited, the curve terminates at point H. Here OH is the load current due to the voltage generated by residual flux.

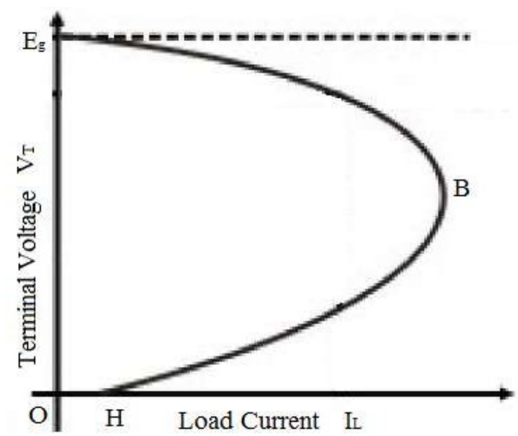


Fig. 1.33 Load characteristics of DC shunt generator

Compound generator

As introduced earlier, compound machines have both series and shunt field coils. Series field coil may be connected in such a way that the mmf produced by it aids the shunt field mmf-then the machine is said to be cumulative compound machine, otherwise if the series field mmf acts in opposition with the shunt field mmf – then the machine is said to be differential compound machine.

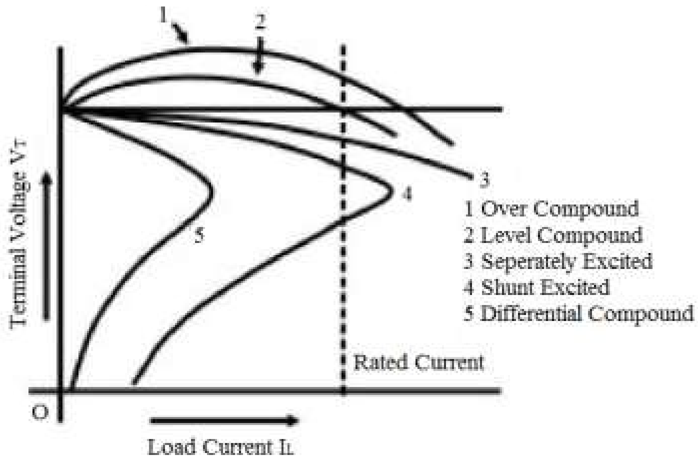


Fig. 1.34 Load characteristics of DC compound generator

In a compound generator, series field coil current is load dependent. Therefore, for a cumulatively compound generator, with the increase of load, flux per pole increases. This in turn increases the generated emf and terminal voltage. Unlike a shunt motor, depending on the strength of the series field mmf, terminal voltage at full load current may be same or more than the no load voltage. When the terminal voltage at rated current is same that at no load condition, then it is called a level compound generator. If however, terminal voltage at rated current is more than the voltage at no load, it is called a over compound generator. The load characteristic of a cumulative compound generator will naturally be above the load characteristic of a shunt generator as depicted in figure 1.34. At load current higher than the rated current, terminal voltage starts decreasing due to saturation, armature reaction effect and more drop in armature and series field resistances.

Parallel Operation of DC Generator

Advantages of DC generator operating in parallel

In a dc power plant, power is usually supplied from several generators of small ratings connected in parallel instead of from one large generator. This is due to the following reasons:

a. *Continuity of service:*

If a single large generator is used in the power plant, then in case of its breakdown, the whole plant will be shut down. However, if power is supplied from a number of small units operating in parallel, then in case of failure of one unit, the continuity of supply can be maintained by other healthy units.

b. ***Efficiency:***

Generators run most efficiently when loaded to their rated capacity. Therefore, when load demand on power plant decreases, one or more generators can be shut down and the remaining units can be efficiently loaded.

c. ***Maintenance and repair:***

Generators generally require routine-maintenance and repair. Therefore, if generators are operated in parallel, the routine or emergency operations can be performed by isolating the affected generator while load is being supplied by other units. This leads to both safety and economy.

d. ***Increasing plant capacity:***

In the modern world of increasing population, the use of electricity is continuously increasing. When added capacity is required, the new unit can be simply paralleled with the old units.

e. ***Non-availability of single large unit:***

In many situations, a single unit of desired large capacity may not be available. In that case a number of smaller units can be operated in parallel to meet the load requirement. Generally a single large unit is more expensive.

Connecting Shunt Generators in Parallel:

The generators in a power plant are connected in parallel through bus-bars. The bus-bars are heavy thick copper bars and they act as +ve and -ve terminals. The positive terminals of the generators are connected to the +ve side of bus-bars and negative terminals to the negative side of bus-bars. Fig. 1.35 shown shunt generator 1 connected to the bus-bars and supplying load. When the load on the power plant increases beyond the capacity of this generator, the second shunt generator 2 is connected in parallel with the first to meet the increased load demand.

The procedure for paralleling generator 2 with generator 1 is as under:

i. The prime mover of generator 2 is brought up to the rated speed. Now switch S2 in the field circuit of the generator 2 is closed.

ii. Next circuit breaker CB-2 is closed and the excitation of generator 2 is adjusted till it generates voltage equal to the bus-bars voltage. This is indicated by voltmeter V2.

iii. Now the generator 2 is ready to be paralleled with generator 1. The main switch DPST2 is closed, thus putting generator 2 in parallel with generator 1. Note that generator 2 is not supplying any load because its generated emf is equal to bus-bars voltage. The generator is said to be “floating” (i.e. not supplying any load) on the bus-bars.

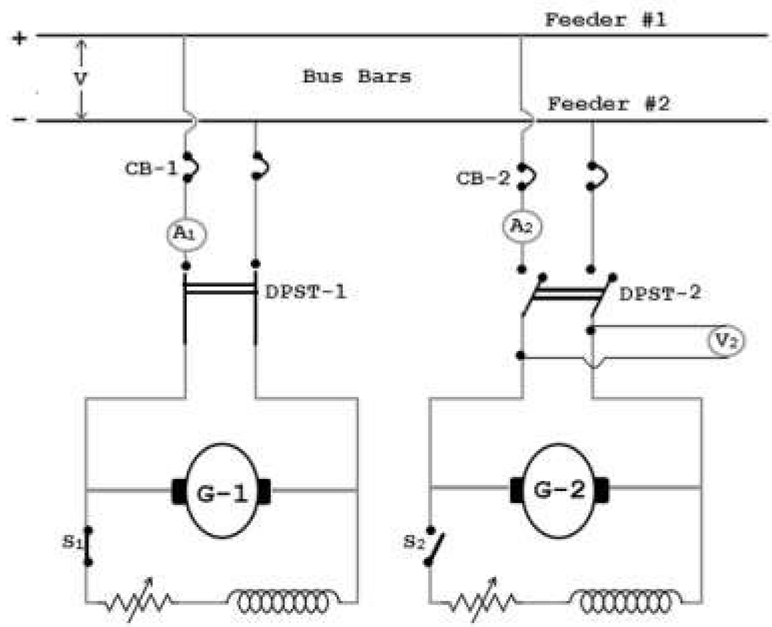


Fig. 1.35 Schematic diagram of DC generator connected in parallel

iv. If generator 2 is to deliver any current, then its generated voltage E_g should be greater than the bus-bars voltage V_T. In that case, current supplied by it is $I = (E_g - V_T)/R_a$ where R_a is the resistance of the armature circuit. By increasing the field current (and hence induced emf E_g), the generator 2 can be made to supply proper amount of load.

v. The load may be shifted from one shunt generator to another merely by adjusting the field excitation. Thus if generator 1 is to be shut down, the whole load can be shifted onto generator 2 provided it has the capacity to supply that load. In that case, reduce the current supplied by generator 1 to zero (This will be indicated by ammeter A1) open C.B.-1 and then open the main switch DPST1.

Equalizer Bar:

Compound Generators in Parallel: Under-compounded generators also operate satisfactorily in parallel but over compounded generators will not operate satisfactorily unless their series fields are paralleled. This is achieved by connecting two negative brushes together as shown in Fig. 1.36. The conductor used to connect these brushes is generally called equalizer bar. Suppose that an attempt is made to operate the two generators in parallel without an equalizer bar. If, for any reason, the current supplied by generator 1 increases slightly, the current in its series field will increase and raise the generated voltage. This will cause generator 1 to take more load. Since total load supplied to the system is constant, the current in generator 2 must decrease and as a result its series field is weakened. Since this effect is cumulative, the generator 1 will take the entire load and drive generator 2 as a motor. After machine 2 changes from a generator to a motor, the current in the shunt field will remain in the same direction, but the current in the armature and series field will reverse. Thus the magnetizing action, of the series field opposes that of the shunt field. As the current taken by the machine 2 increases, the demagnetizing action of series field becomes greater and the resultant field becomes weaker. The resultant field will finally become zero and at that time machine 2 will be short circuited machine 1, opening the breaker of either or both machines.

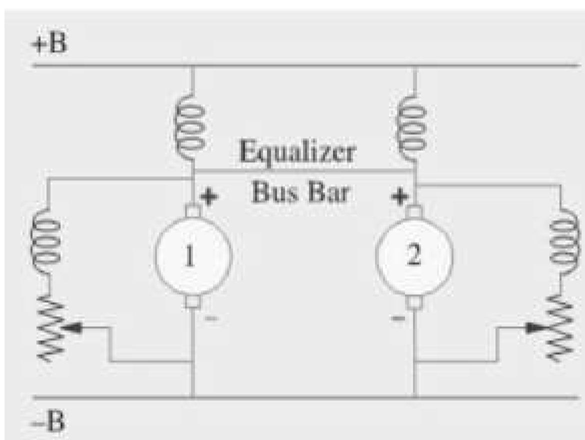


Fig. 1.36 Connection of equalizer bar in parallel connection of DC compound generator

When the equalizer bar is used, a stabilizing action exists and neither machine tends to take all the load.To consider this, suppose that current delivered by generator 1 increases. The increased current will not only pass through the series field of generator 1 but also through the equalizer bar and series field of generator 2. Therefore, the voltage of both the machines increases and the generator 2 will take a part of the load.

Load Sharing:

The load sharing between shunt generators in parallel can be easily regulated because of their drooping characteristics. The load may be shifted from one generator to another merely by adjusting the field excitation. Let us discuss the load sharing of two generators which have unequal no-load voltages. Let E_1, E_2 = no-load voltages of the two generators R_1, R_2 = their armature resistances

V_T = common terminal voltage (Bus-bars voltage). Then

$$I_1=\frac{E_1\text{-}V_T}{R_1} \text{ and } I_2=\frac{E_2\text{-}V_T}{R_2}$$

Thus current output of the generators depends upon the values of E_1 and E_2. These values may be changed by field rheostats. The common terminal voltage (or bus-bars voltage) will depend upon (i) the emfs of individual generators and (ii) the total load current supplied. It is generally desired to keep the busbars voltage constant. This can be achieved by adjusting the field excitations of the generators operating in parallel.

Transformers

2.1 Introduction

The transformer is a device that transfers electrical energy from one electrical circuit to another electrical circuit. The two circuits may be operating at different voltage levels but always work at the same frequency. Basically transformer is an electro-magnetic energy conversion device. It is commonly used in electrical power system and distribution systems. It can change the magnitude of alternating voltage or current from one value to another. This useful property of transformer is mainly responsible for the widespread use of alternating currents rather than direct currents i.e., electric power is generated, transmitted and distributed in the form of alternating current. Transformers have no moving parts, rugged and durable in construction, thus requiring very little attention. They also have a very high efficiency as high as 99%.

2.2. Single Phase Transformer

A transformer is a static device of equipment used either for raising or lowering the voltage of an a.c. supply with a corresponding decrease or increase in current. It essentially consists of two windings, the primary and secondary, wound on a common laminated magnetic core as shown in Fig 1. The winding connected to the a.c. source is called primary winding (or primary) and the one connected to load is called secondary winding (or secondary). The alternating voltage V_1 whose magnitude is to be changed is applied to the primary.

Depending upon the number of turns of the primary (N_1) and secondary (N_2), an alternating e.m.f. E_2 is induced in the secondary. This induced e.m.f. E_2 in the secondary causes a secondary current I_2. Consequently, terminal voltage V_2 will appear across the load.

If $V_2 > V_1$, it is called a step up-transformer.

If $V_2 < V_1$, it is called a step-down transformer.

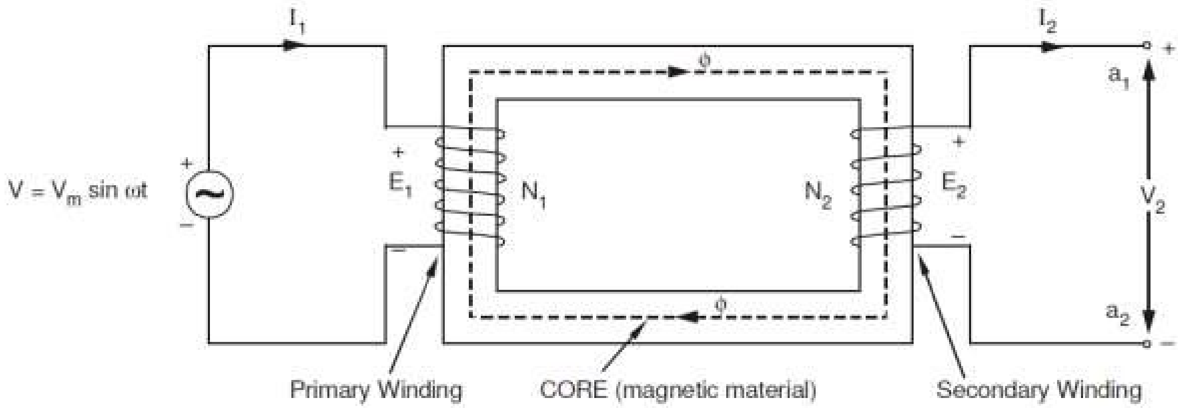


Fig. 2.1 Schematic diagram of single phase transformer

Constructional Details

Depending upon the manner in which the primary and secondary windings are placed on the core, and the shape of the core, there are two types of transformers, called (a) core type, and (b) shell type.

Core-type and Shell-type Construction

In core type transformers, the windings are placed in the form of concentric cylindrical coils placed around the vertical limbs of the core. The low-voltage (LV) as well as the high-voltage (HV) winding are made in two halves, and placed on the two limbs of core. The LV winding is placed next to the core for economy in insulation cost. Figure 2.1(a) shows the cross-section of the arrangement. In the shell type transformer, the primary and secondary windings are wound over the central limb of a three-limb core as shown in Figure 2.1(b). The HV and LV windings are split into a number of sections, and the sections are interleaved or sandwiched i.e. the sections of the HV and LV windings are placed alternately.

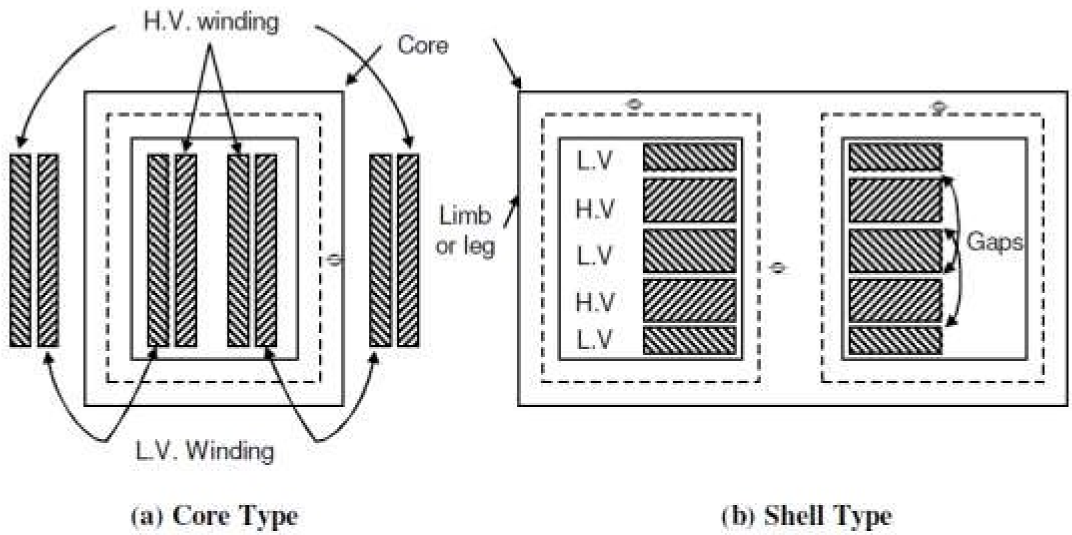


Fig: 2.1 Core type & shell type transformer

Core

The core is built-up of thin steel laminations insulated from each other. This helps in reducing the eddy current losses in the core, and also helps in construction of the transformer. The steel used for core is of high silicon content, sometimes heat treated to produce a high permeability and low hysteresis loss. The material commonly used for core is CRGO (Cold Rolled Grain Oriented) steel. Conductor material used for windings is mostly copper. However, for small distribution transformer aluminium is also sometimes used. The conductors, core and whole windings are insulated using various insulating materials depending upon the voltage.

Insulating Oil

In oil-immersed transformer, the iron core together with windings is immersed in insulating oil. The insulating oil provides better insulation, protects insulation from moisture and transfers the heat produced in core and windings to the atmosphere.

The transformer oil should possess the following qualities:

(a) High dielectric strength,

(b) Low viscosity and high purity,

(c) High flash point, and

(d) Free from sludge.

Transformer oil is generally a mineral oil obtained by fractional distillation of crude oil.

Tank and Conservator

The transformer tank contains core wound with windings and the insulating oil. In large transformers small expansion tank is also connected with main tank is known as conservator. Conservator provides space when insulating oil expands due to heating. The transformer tank is provided with tubes on the outside, to permits circulation of oil, which aides in cooling. Some additional devices like breather and Buchholz relay are connected with main tank. Buchholz relay is placed between main tank and conservator. It protect the transformer under extreme heating of transformer winding. Breather protects the insulating oil from moisture when the cool transformer sucks air inside. The silica gel filled breather absorbs moisture when air enters the tank. Some other necessary parts are connected with main tank like, Bushings, Cable Boxes, Temperature gauge, Oil gauge, Tappings, etc.

Principle of Operation

When an alternating voltage V1 is applied to the primary, an alternating flux ϕ is set up in the core. This alternating flux links both the windings and induces e.m.f.s E1 and E2 in them according to Faraday's laws of electromagnetic induction. The e.m.f. E1 is termed as primary e.m.f. and e.m.f. E2 is termed as secondary e.m.f.

Clearly, $$E_1 = -N_1 \frac{d\phi}{dt}$$

and $$E_2 = -N_2 \frac{d\phi}{dt}$$

$$\therefore \quad \frac{E_2}{E_1} = \frac{N_2}{N_1}$$

Note that magnitudes of E2 and E1 depend upon the number of turns on the secondary and primary respectively.

If $N_2 > N_1$, then $E_2 > E_1$ (or $V_2 > V_1$) and we get a step-up transformer. If $N_2 < N_1$, then $E_2 < E_1$ (or $V_2 < V_1$) and we get a step-down transformer.

If load is connected across the secondary winding, the secondary e.m.f. E_2 will cause a current I_2 to flow through the load. Thus, a transformer enables us to transfer a.c. power from one circuit to another with a change in voltage level.

The following points may be noted carefully:

(a) The transformer action is based on the laws of electromagnetic induction.

(b) There is no electrical connection between the primary and secondary.

(c) The a.c. power is transferred from primary to secondary through magnetic flux.

(d) There is no change in frequency i.e., output power has the same frequency as the input power.

(e) The losses that occur in a transformer are:

(a) ***core losses***—eddy current and hysteresis losses

(b) ***copper losses***—in the resistance of the windings

In practice, these losses are very small so that output power is nearly equal to the input primary power. In other words, a transformer has very high efficiency.

E.M.F. Equation of a Transformer

Consider that an alternating voltage V_1 of frequency f is applied to the primary as shown in Fig.2.3. The sinusoidal flux ϕ produced by the primary can be represented as:

$\phi = \phi_m \sin\omega t$

When the primary winding is excited by an alternating voltage V_1, it is circulating alternating current, producing an alternating flux ϕ.

ϕ - Flux

ϕ_m - maximum value of flux

N_1 - Number of primary turns

N_2 - Number of secondary turns

f - Frequency of the supply voltage

E_1 - R.M.S. value of the primary induced e.m.f

E_2 - R.M.S. value of the secondary induced e.m.f

The instantaneous e.m.f. e_1 induced in the primary is -

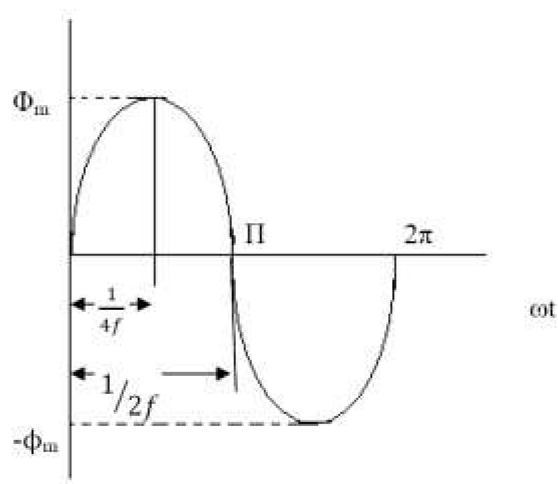


Fig. 2.3

From Faraday's law of electromagnetic induction -

$$\text{Average e.m.f per turns} = \frac{d\emptyset}{dt}$$

$d\phi$ = change in flux
dt = time required for change in flux

The flux increases from zero value to maximum value ϕm in 1/4f of the time period that is in 1/4f seconds.

The change of flux that takes place in 1/4f seconds = $\phi_m - 0 = \phi_m$ webers

$$\frac{d\phi}{dt} = \frac{dt}{1/4f} = 4f\phi_m\ w_b/sec.$$

Since flux ϕ varies sinusoidally, the R.m.s value of the induced e.m.f is obtained by multiplying the average value with the form factor

$$\text{Form factor of a sinwave} = \frac{\text{R.m.s value}}{\text{Average value}} = 1.11$$

R.M.S Value of e.m.f induced in one turns = $4\phi_m f \times 1.11$ Volts.

= $4.44\phi_m f$ Volts.

R.M.S Value of e.m.f induced in primary winding = $4.44\phi_m f$ N1Volts.

R.M.S Value of e.m.f induced in secondary winding = $4.44\phi_m f$ N2Volts.

The expression of E_1 and E_2 are called e.m.f equation of a transformer

$$\boxed{\begin{array}{l} V_1 = E_1 = 4.44\phi_m f N_1 \text{Volts.} \\ V_2 = E_2 = 4.44\phi_m f N_2 \text{Volts.} \end{array}}$$

Voltage Ratio

Voltage transformation ratio is the ratio of e.m.f induced in the secondary winding to the e.m.f induced in the primary winding.

$$\frac{E2}{E1} = \frac{4.44\phi mf\ N2}{4.44\phi mf\ N1}$$

$$\boxed{\frac{E2}{E1} = \frac{N2}{N1} = K}$$

This ratio of secondary induced e.m.f to primary induced e.m.f is known as voltage transformation ratio

$$E_2 = KE_1 \qquad \text{where } K = \frac{N2}{N1}$$

1. If $N_2>N_1$ i.e. K>1 we get $E_2>E_1$ then the transformer is called step up transformer.

2. If $N_2< N_1$ i.e. K<1 we get $E_2< E_2$ then the transformer is called step down transformer.

3. If $N_2= N_1$ i.e. K=1 we get $E_2= E_2$ then the transformer is called isolation transformer or 1:1

transformer.

Current Ratio

Current ratio is the ratio of current flow through the primary winding (I_1) to the current flowing through the secondary winding (I_2). In an ideal transformer -

Apparent input power = Apparent output power.

$$V_1 I_1 = V_2 I_2$$

$$\frac{I1}{I2} = \frac{V2}{V1} = \frac{N2}{N1} = K$$

Volt-Ampere Rating

i) The transformer rating is specified as the products of voltage and current (VA rating).

ii) On both sides, primary and secondary VA rating remains same. This rating is generally expressed in KVA (Kilo Volts Amperes rating).

$$\frac{V1}{V2} = \frac{I2}{I1} = K$$

$$V_1 I_1 = V_2 I_2$$

$$\text{KVA Rating of a transformer} = \frac{V1I1}{1000} = \frac{V2I2}{1000} \quad \text{(1000 is to convert KVA to VA)}$$

V_1 and V_2 are the V_t of primary and secondary by using KVA rating we can calculate I_1 and I_2 Full load current and it is safe maximum current.

$$I_1 \text{ Full load current} = \frac{KVA\ Rating\ X\ 1000}{V1}$$

$$I_1 \text{ Full load current} = \frac{KVA\ Rating\ X\ 1000}{V2}$$

Transformer on No-load

a) Ideal transformer

b) Practical transformer

a) Ideal Transformer

An ideal transformer is one that has

(i) No winding resistance

(ii) No leakage flux i.e., the same flux links both the windings

(iii) No iron losses (i.e., eddy current and hysteresis losses) in the core

Although ideal transformer cannot be physically realized, yet its study provides a very powerful tool in the analysis of a practical transformer. In fact, practical transformers have properties that approach very close to an ideal transformer.

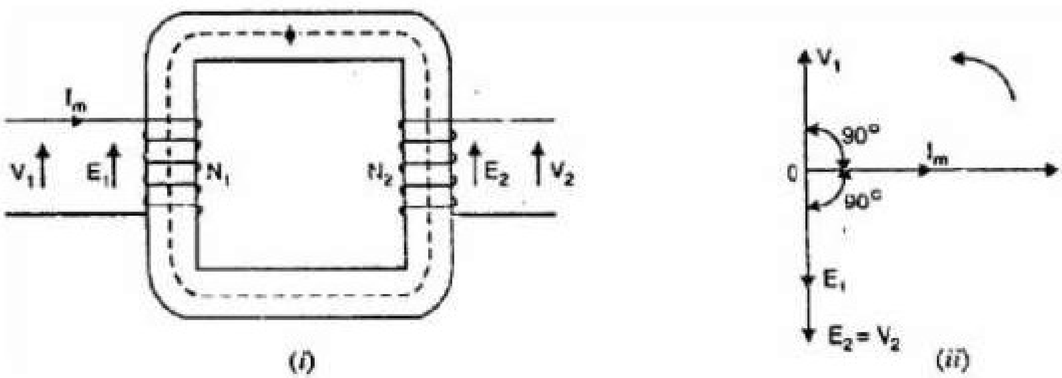


Fig: 2.4

Consider an ideal transformer on no load i.e., secondary is open-circuited as shown in *Fig.2.4 (i)*. under such conditions, the primary is simply a coil of pure inductance. When an alternating voltage V_1 is applied to the primary, it draws a small magnetizing current Im which lags behind the applied voltage by 90°. This alternating current Im produces an alternating flux ϕ which is proportional to and in phase with it. The alternating flux ϕ links both the windings and induces e.m.f. E1 in the primary and e.m.f. E_2 in the secondary. The primary e.m.f. E1 is, at every instant, equal to and in opposition to V_1 (Lenz's law). Both e.m.f.s E_1 and E2 lag behind flux ϕ by 90°.However, their magnitudes depend upon the number of primary and secondary turns. *Fig. 2.4 (ii)* shows the phasor diagram of an ideal transformer on no load. Since flux ϕ is common to both the windings, it has been taken as the reference phasor. The primary e.m.f. E1 and secondary e.m.f. E2 lag behind the flux ϕ by 90°. Note that E_1 and E_2 are in phase. But E1 is equal to V1 and 180° out of phase with it.

$$\boxed{\frac{E2}{E1} = \frac{V2}{V1} = K}$$

Phasor Diagram

i) Φ (flux) is reference

ii) Im produce ϕ and it is in phase with ϕ, V_1 Leads Im by 90˚

iii) E1 and E2 are in phase and both opposing supply voltage V1, winding is purely inductive

So current has to lag voltage by 90˚.

iv) The power input to the transformer

$$P = V_1 . I_1 . \cos(90^\circ) \ldots\ldots\ldots\ldots\ldots (\cos 90^\circ = 0)$$

$$P = 0 \text{ (ideal transformer)}$$

***b)i) Practical Transformer* on no load**

A practical transformer differs from the ideal transformer in many respects. The practical transformer has (i) iron losses (ii) winding resistances and (iii) Magnetic leakage

(i) Iron losses. Since the iron core is subjected to alternating flux, there occurs eddy current and hysteresis loss in it. These two losses together are known as iron losses or core losses. The iron losses depend upon the supply frequency, maximum flux density in the core, volume of the core etc. It may be noted that magnitude of iron losses is quite small in a practical transformer.

(ii) Winding resistances. Since the windings consist of copper conductors, it immediately follows that both primary and secondary will have winding resistance. The primary resistance R_1 and secondary resistance R_2 act in series with the respective windings as shown in Fig. When current flows through the windings, there will be power loss as well as a loss in voltage due to IR drop. This will affect the power factor and E_1 will be less than V_1 while V_2 will be less than E_2.

Consider a practical transformer on no load i.e., secondary on open-circuit as Shown in Fig 2.5.

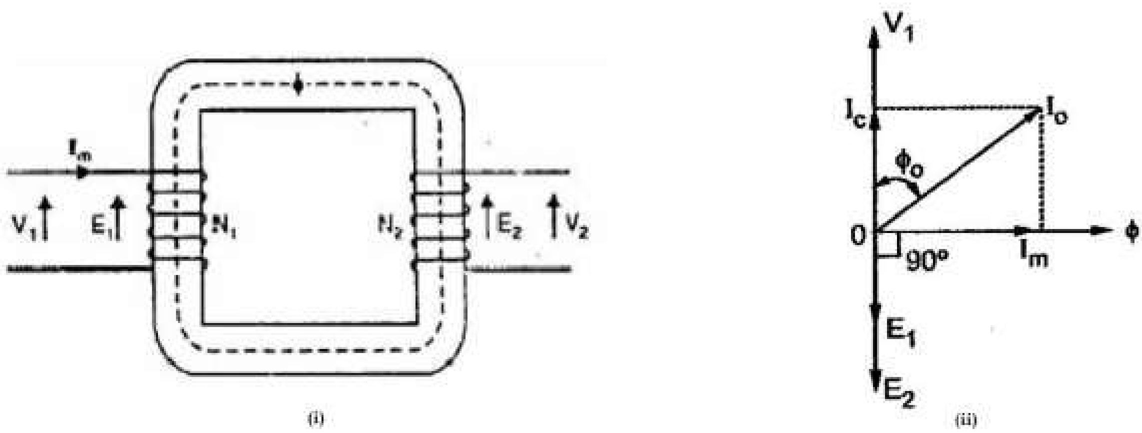


Fig: 2.5 Phasor diagram of transformer at noload

Here the primary will draw a small current I_0 to supply -

(i) the iron losses and

(ii) a very small amount of copper loss in the primary.

Hence the primary no load current I0 is not 90° behind the applied voltage V_1 but lags it by an angle ϕ_0 < 90° as shown in the phasor diagram.

No load input power, $W_0 = V_1 I_0 \cos \phi 0$

As seen from the phasor diagram in Fig.2.5 (ii), the no-load primary current I_0

(i) The component I_c in phase with the applied voltage V_1. This is known as active or working or iron loss component and supplies the iron loss and a very small primary copper loss.

$I_c = I_0 \cos \phi_0$

The component Im lagging behind V_1 by 90° and is known as magnetizing component. It is this component which produces the mutual flux ϕ in the core.

$I_m = I_0 \sin \phi_0$

Clearly, Io is phasor sum of Im and I_c,

$$I_o = \sqrt{Im^2 + Ic^2}$$

$$\text{No load P.F.,} \quad \cos\phi o = \frac{Ic}{Io}$$

The no load primary copper loss (i.e. Io^2R1) is very small and may be neglected.

Therefore, the no load primary input power is practically equal to the iron loss in the transformer i.e.,

No load input power, $W_0 = V_1 I_o \cos\phi_o = Pi$ = Iron loss

***b) ii)Practical Transformer* on Load**

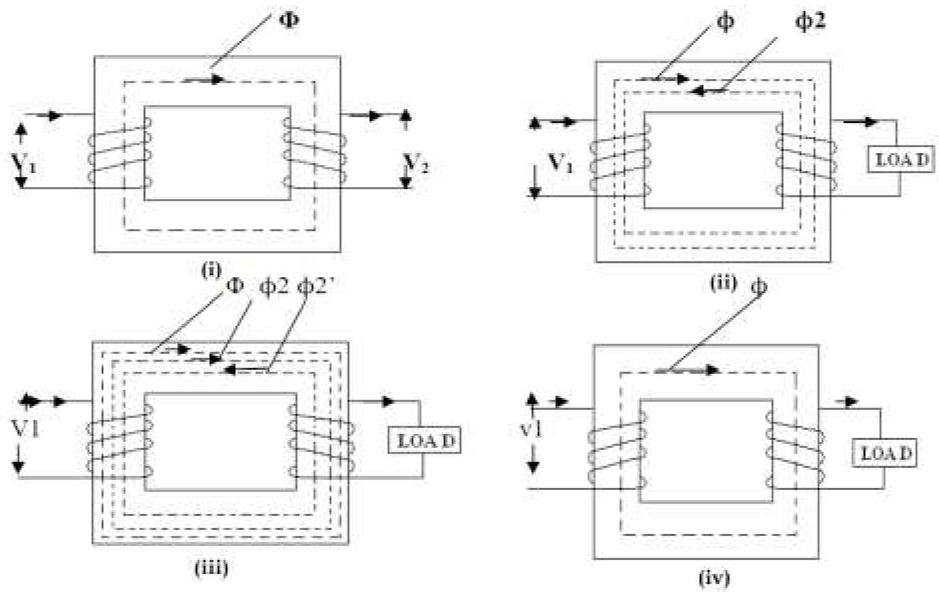


Fig: 2.6

At no load, there is no current in the secondary so that $V_2 = E_2$. On the primary side, the drops in R_1 and X_1, due to I_0 are also very small because of the smallness of I_0. Hence, we can say that at no load, $V_1 = E_1$.

i) When transformer is loaded, the secondary current I_2 is flows through the secondary winding.

ii) Already I_m magnetizing current flow in the primary winding fig. 2.6(i).

iii) The magnitude and phase of I_2 with respect to V_2 is determined by the characteristics of the load.

a) I_2 in phase with V_2 (resistive load)

b) I_2 lags with V_2 (Inductive load)

c) I_2 leads with V_2 (capacitive load)

iv) Flow of secondary current I_2 produce new Flux ϕ_2 fig.2.6 (ii)

v) Φis main flux which is produced by the primary to maintain the transformer as constant magnetising component.

vi) Φ_2 opposes the main flux ϕ, the total flux in the core reduced. It is called demagnetising Ampere-turns due to this E_1 reduced.

vii) To maintain the ϕ constant primary winding draws more current (I_2') from the supply (load component of primary) and produce ϕ_2' flux which is oppose ϕ_2 (but in same direction as ϕ), to maintain flux constant flux constant in the core fig.2.6 (iii).

viii) The load component current I_2' always neutralizes the changes in the load.

ix) Whatever the load conditions, the net flux passing through the core is approximately the same as at no-load. An important deduction is that due to the constancy of core flux at all loads, the core loss is also practically the same under all load conditions fig.2.6 (iv).

$$\boxed{\Phi_2 = \phi_2' \quad N_2I_2 = N_1I_2' \quad I_2' = \frac{N2}{N1} X I_2 = KI_2}$$

Phasor Diagram

i) Take (ϕ) flux as reference for all load

ii) The no load I_o which lags by an angle ϕo, $Io = \sqrt{Ic^2 + Im^2}$.

iii) The load component I_2', which is in anti-phase with I_2 and phase of I_2 is decided by the load.

iv) Primary current I_1 is vector sum of I_o and I_2'

$$\vec{I_1} - \vec{I_o} + \vec{I_2'}$$

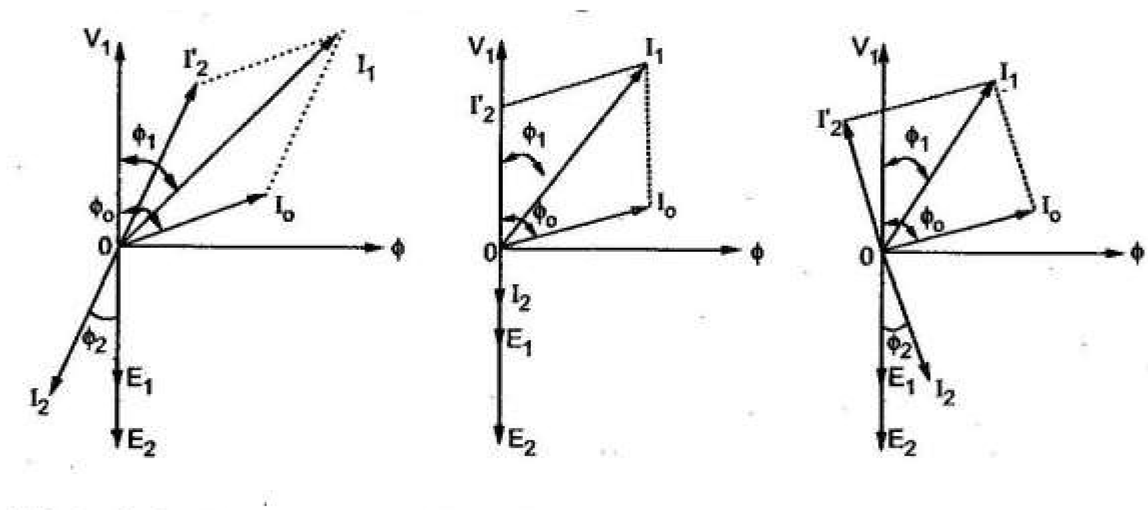
$$I_1 = \sqrt{Io^2 + I2'^2}$$

a) If load is Inductive, I_2 lags E_2 by ϕ_2, shown in phasor diagram fig 2.7 (a).

b) If load is resistive, I_2 in phase with E_2 shown in phasor diagram fig. 2.7 (b).

c) If load is capacitive load, I_2 leads E_2 by ϕ_2 shown in phasor diagram fig. 2.7 (c).

For easy understanding at this stage here we assumed E_2 is equal to V_2 neglecting various drops.



(a) Inductive load (b) Resistive load (c) Capacitive load

Fig: 2.7.a

$$I_1 \cong I_2'$$

Balancing the ampere – turns

$$N_1 I_2' = N_1 I_1 + N_2 I_2$$

$$\frac{I1}{I2} = \frac{N2}{N1} = K$$

Now we going to construct complete phasor diagram of a transformer (shown in Fig: 2.7.b)

Effect of Winding Resistance

In practical transformer it process its own winding resistance causes power loss and also the voltage drop.

R_1 – primary winding resistance in ohms.
R_2 – secondary winding resistance in ohms.

The current flow in primary winding make voltage drop across it is denoted as I_1R_1 here supply voltage V_1 has to supply this drop primary induced e.m.f E_1 is the vector difference between V_1 and I_1R_1.

$$\vec{E_1} = \vec{V_1} - \vec{I_1R_1}$$

Similarly the induced e.m.f in secondary E_2, The flow of current in secondary winding makes voltage drop across it and it is denoted as I_2R_2 here E_2 has to supply this drop.

The vector difference between E_2 and I_2R_2

$$\vec{V_2} = \vec{E_2} - \vec{I_2R_2}$$ (Assuming as purely resistive drop here.)

Equivalent Resistance

1) It would now be shown that the resistances of the two windings can be transferred to any one of the two winding.

2) The advantage of concentrating both the resistances in one winding is that it makes calculations very simple and easy because one has then to work in one winding only.

3) Transfer to any one side either primary or secondary without affecting the performance of the transformer.

The total copper loss due to both the resistances.

$$\text{Total copper loss} = I_1^2R_1 + I_2^2R_2$$
$$= I_1^2[R_1 + \frac{I2^2}{I1^2}]$$
$$= I_1^2[R_1 + \frac{1}{K} R_2]$$

$\frac{R2}{K^2}$ is the resistance value of R_2 shifted to primary side and denoted as R_2'.

R_2' is the equivalent resistance of secondary referred to primary

$$\boxed{R_2' = \frac{R2}{K^2}}$$

Equivalent resistance of transformer referred to primary fig (ii)

$$\boxed{R_{1e} = R_1 + R_2' = R_1 + \frac{R2}{K^2}}$$

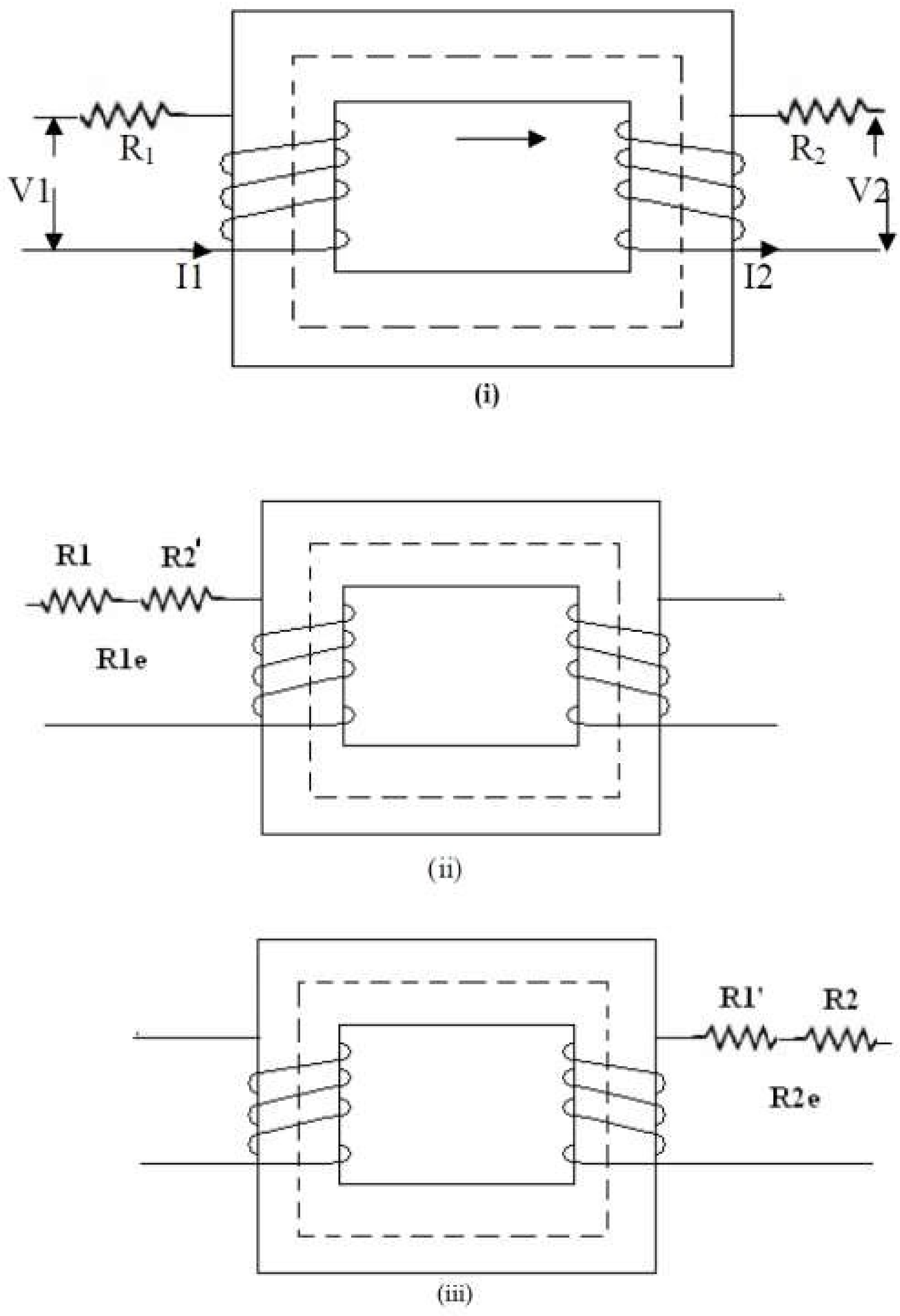


Fig:2.8

Similarly it is possible to refer the equivalent resistance to secondary winding.

Total copper loss = $I_1^2R_1 + I_2^2R_2 = I_2^2 [\frac{I1^2}{I2^2} R_1 + R_2]$

$= I_2^2[K^2R_1 + R_2]$

K^2R1 is primary resistance referred to secondary denoted as R1'.

$R_1' = K^2R_1$

Equivalent resistance of transformer referred to secondary, denoted as R2e

$$\boxed{R_{2e} = R_2 + R_1' = R_2 + K^2R_1}$$

Total copper loss = $I_2^2 R_{2e}$

Note:

Note:

i) When a resistance is to be transferred from the primary to secondary, it must be multiplied by K^2, it must be divided by K^2 while transferred from the secondary to primary.

High voltage side ➡ low current side ➡ high resistance side

Low voltage side ➡ high current side ➡ low resistance side

Effect of Leakage Reactance

i) It has been assumed that all the flux linked with primary winding also links the secondary winding. But, in practice, it is impossible to realize this condition.

ii) However, primary current would produce flux ϕ which would not link the secondary winding. Similarly, current would produce some flux ϕ that would not link the primary winding.

iii) The flux ϕ_{L1} complete its magnetic circuit by passing through air rather than around the core, as shown in fig.2.9. This flux is known as primary leakage flux and is proportional to the primary ampere – turns alone because the secondary turns do not links the magnetic circuit of ϕ_{L1}. It induces an e.m.f e_{L1} in primary but not in secondary.

iv) The flux ϕ_{L2} complete its magnetic circuit by passing through air rather than around the core, as shown in fig. This flux is known as secondary leakage flux and is proportional to the secondary ampere – turns alone because the primary turns do not links the magnetic circuit of ϕ_{L2}. It induces an e.m.f e_{L2} in secondary but not in primary.

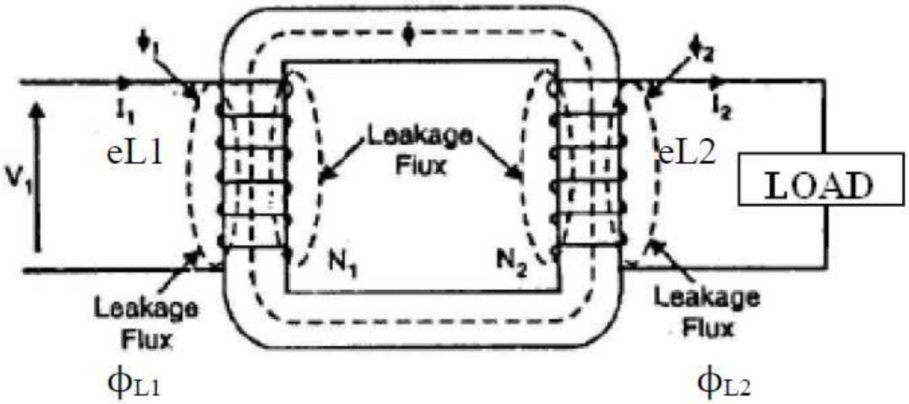


Fig: 2.9

ϕ_{L1} – primary leakage flux
ϕ_{L2} – secondary leakage flux
e_{L1} – self induced e.m.f (primary)
e_{L2} –self induced e.m.f (secondary)

Equivalent Leakage Reactance

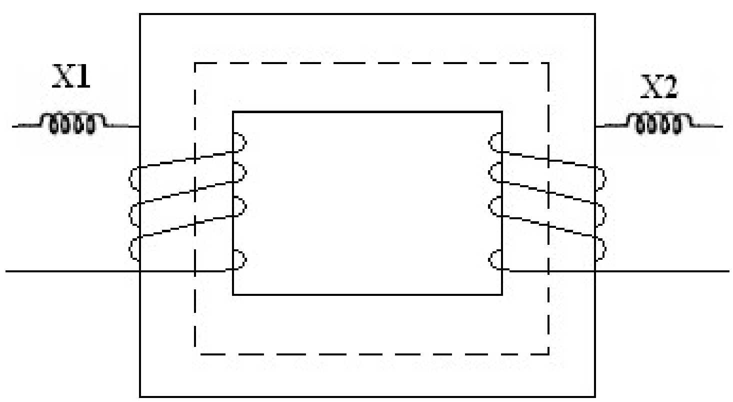


Fig: 2.10

Similarly to the resistance, the leakage reactance also can be transferred from primary to secondary. The relation through K^2 remains same for the transfer of reactance as it is studied earlier for the resistance

X_1 – leakage reactance of primary.

X_2 - leakage reactance of secondary.

Then the total leakage reactance referred to primary is X_{1e} given by

$$X_{1e} = X_1 + X_2'$$

$$X_2' = \frac{X2}{K^2}$$

The total leakage reactance referred to secondary is X_{2e} given by

$$X_{2e} = X_2 + X_1'$$
$$X_1' = K^2 X_1$$

$$\boxed{\begin{aligned} X_{1e} &= X_1 + X_2' \\ X_{2e} &= X_2 + X_1' \end{aligned}}$$

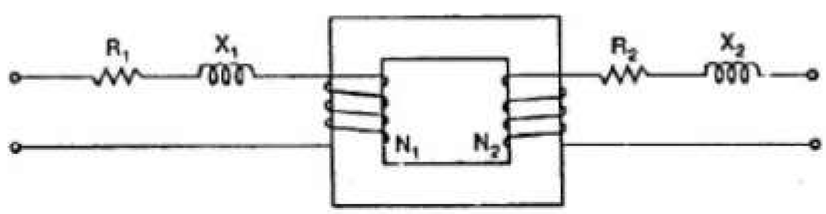
Equivalent Impedance

The transformer winding has both resistance and reactance (R_1, R_2, X_1,X_2).Thus we can say that the total impedance of primary winding isZ_1 which is,

$Z_1 = R_1 + jX_1$ ohms

On secondary winding,

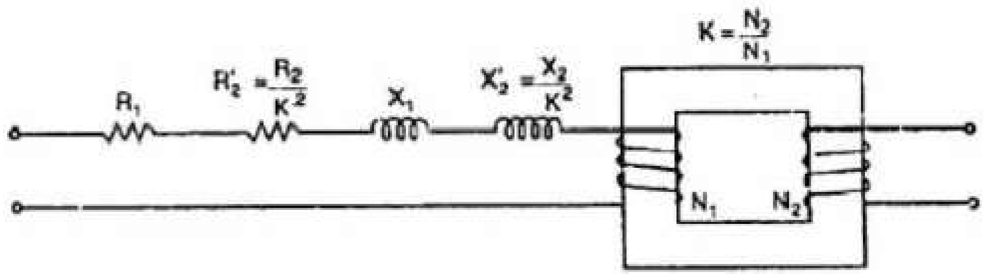
$Z_2 = R_2 + jX_2$ ohms



Individual magnitude of Z1 and Z2 are

$$Z_1 = \sqrt{R1^2 + X1^2}$$
$$Z_2 = \sqrt{R2^2 + X2^2}$$

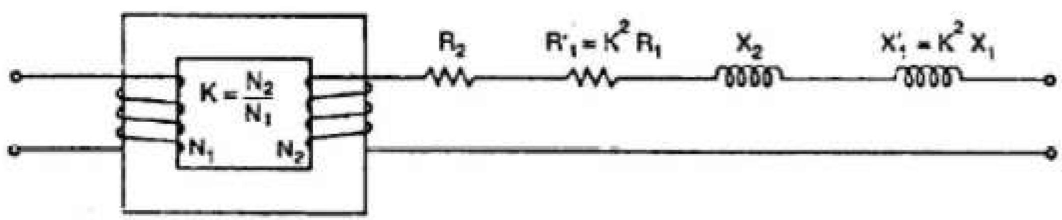
Similar to resistance and reactance, the impedance also can be referred to any one side,



Z_{1e} = total equivalent impedance referred to primary

$$Z_{1e} = R_{1e} + jX_{1e} = Z_1 + Z_2' = Z_1 + \frac{Z2}{K^2}$$

Z_{2e} = total equivalent impedance referred to secondary.



$$Z_{2e} = R_{2e} + jX_{2e} = Z_2 + Z_1' = Z_2 + K^2 Z_1$$

The magnitudes of Z1e and Z2e

$$Z_1 = \sqrt{R1e^2 + X1e^2}$$
$$Z_2 = \sqrt{R2e^2 + X2e^2}$$

It can be noted that

$Z_{2e} = K^2 Z_{1e}$ and $Z_{1e} = \frac{Z2e}{K^2}$

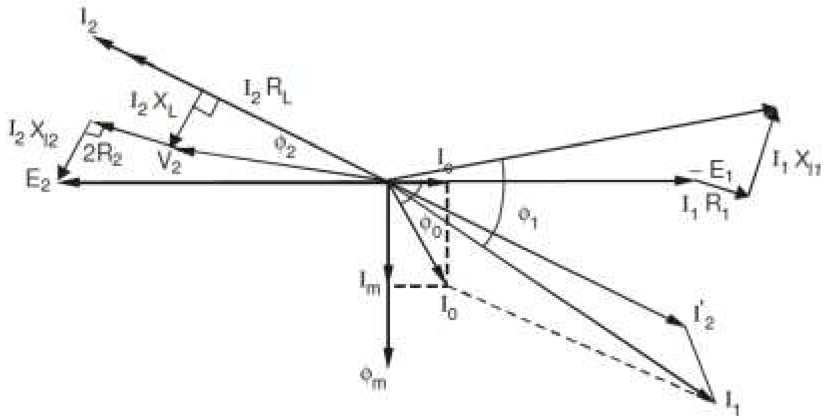
<u>Complete Phasor Diagram of a Transformer *(for Inductive Load or Lagging pf)*</u>

We now restrict ourselves to the more commonly occurring load i.e. inductive along with resistance, which has a lagging power factor.

For drawing this diagram, we must remember that

$$\overline{V_2} = \overline{E}_2 - \overline{I}_2\,(R_2 + j\,X_{l2})$$

and $$\overline{V_1} = -\,\overline{E}_1 + \overline{I}_2\,(R_1 + j\,X_{l1})$$



Equivalent Circuit of Transformer

No load equivalent circuit

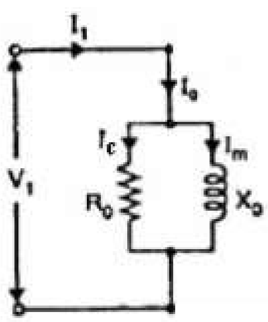


Fig:11

$I_m = I_o \sin \phi_o$ = magnetizing component
$I_c = I_o \cos \phi_o$ = Active component

$R_o = \frac{V1}{Ic}$, $X_o = \frac{V1}{Im}$

i) I_m produces the flux and is assumed to flow through reactance X_o called no load reactance while I_c is active component representing core losses hence is assumed to flow through the resistance R_0

ii) Equivalent resistance is shown in fig.2.12.

iii) When the load is connected to the transformer then secondary current I_2 flows causes voltage drop across R_2 and X_2. Due to I_2, primary draws an additional current.

$$I_2' = \frac{I2}{K}$$

I_1 is the phasor addition of I_o and I_2'. This I_1 causes the voltage drop across primary resistance R_1 and reactance X_1.

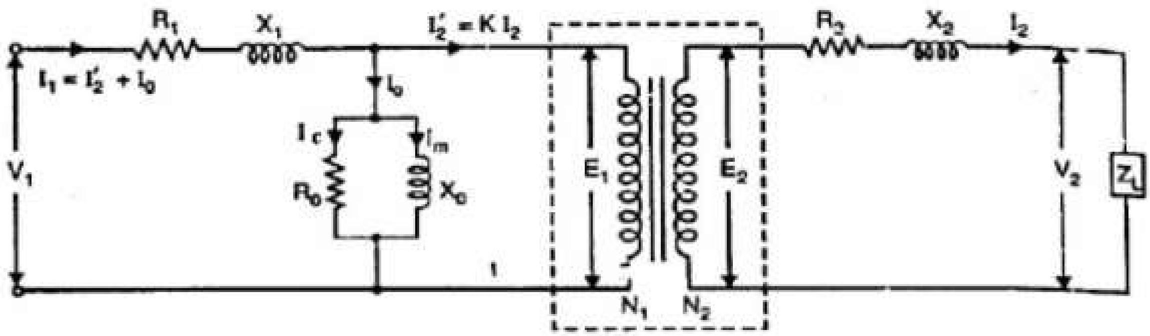


Fig: 2.12

To simplified the circuit the winding is not taken in equivalent circuit while transfer to one side.

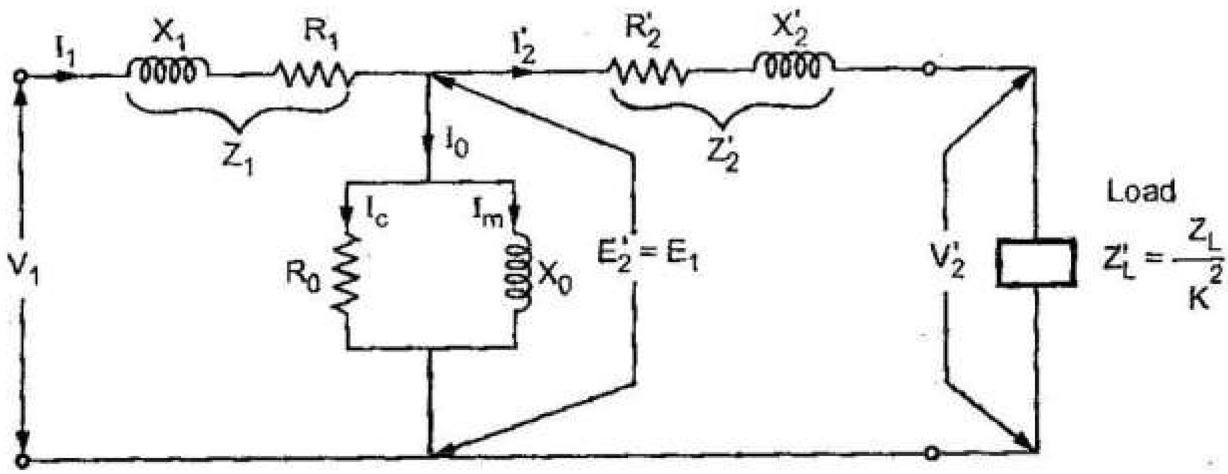


Fig: 2.13

Exact equivalent circuit referred to primary

Transferring secondary parameter to primary -

$$\boxed{R_2' = \frac{R2}{K^2},\ X_2' = \frac{X2}{K^2},\ Z_2' = \frac{Z2}{K^2},\ E_2' = \frac{E2}{K'},\ I_2' = KI_2,\ K = \frac{N2}{N1}}$$

High voltage winding ➡ low current ➡ high impedance

Low voltage winding ➡ high current ➡ low impedance

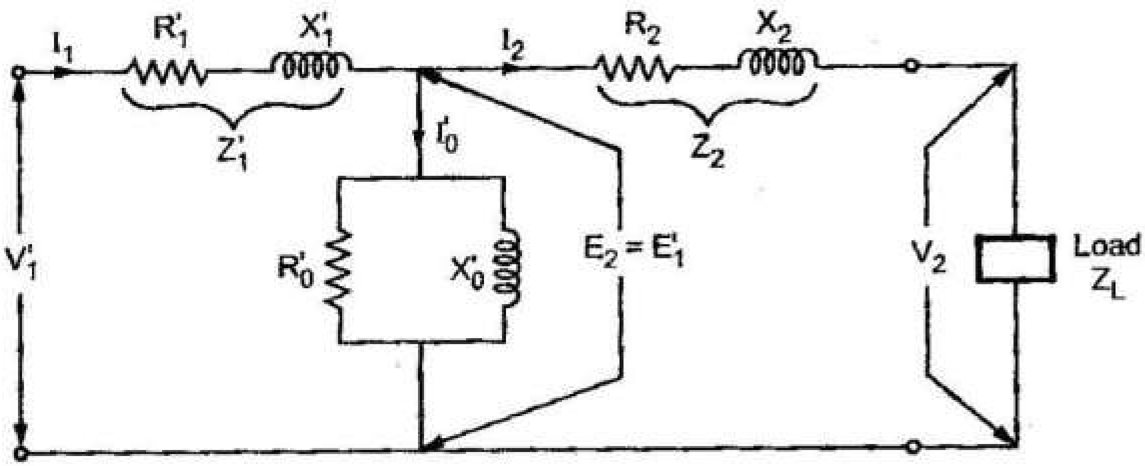


Fig: 2.14

Exact equivalent circuit referred to secondary

$$R_1' = R_1K^2,\ X_1' = K^2X_1\ ,\ E_1' = KE_1$$

$$Z_1' = K^2Z1\ ,\ I_1' = \frac{I1}{K'},\ I_o = \frac{Io}{K}$$

Now as long as no load branch i.e. exciting branch is in between Z_1 and Z_2', the impedances cannot be combined. So further simplification of the circuit can be done. Such circuit is called approximate equivalent circuit.

Approximate Equivalent Circuit

i) To get approximate equivalent circuit, shift the no load branch containing R_o and X_o to the left of R_1 and X_1.

ii) By doing this we are creating an error that the drop across R_1 and X_1 to I_o is neglected due to this circuit because simpler.

iii) This equivalent circuit is called approximate equivalent circuit Fig: 2.15 & Fig: 2.16.

In this circuit new R_1 and R_2' can be combined to get equivalent circuit referred to primary R_{1e},similarly

X_1 and X_2’ can be combined to get X_{1e}.

$$R_{1e} = R_1 + R_2' = R_1 + \frac{R2}{K^2}$$
$$X_{1e} = X_1 + X_2' = X_1 + \frac{X2}{K^2}$$
$$Z_{1e} = R_{1e} + jX_1E, \qquad R_o = \frac{V1}{Ic}, \quad \text{and } X_o = \frac{V1}{Im}$$
$$I_c = I_o \cos\phi_o, \quad \text{and } I_m = I_o \sin\phi_o$$

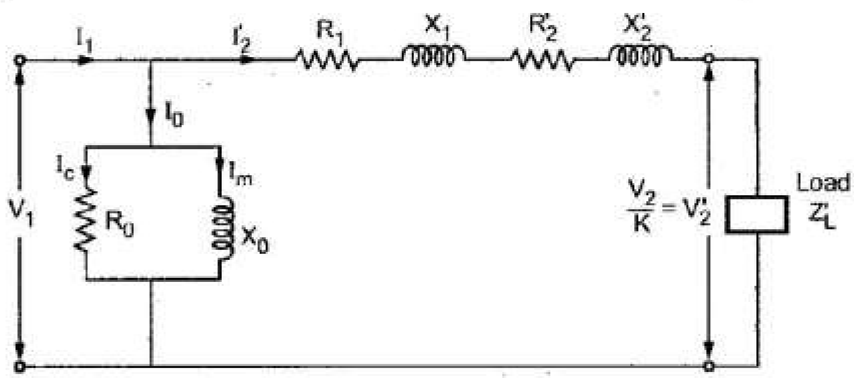


Fig:2.15Approximate equivalent circuit referred to primary

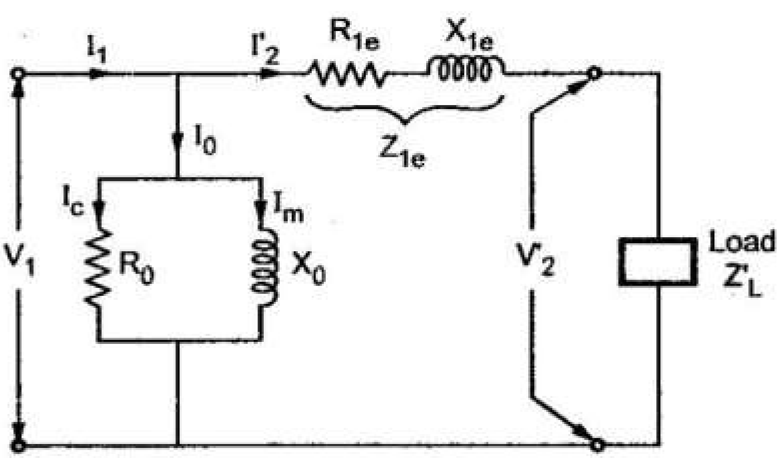
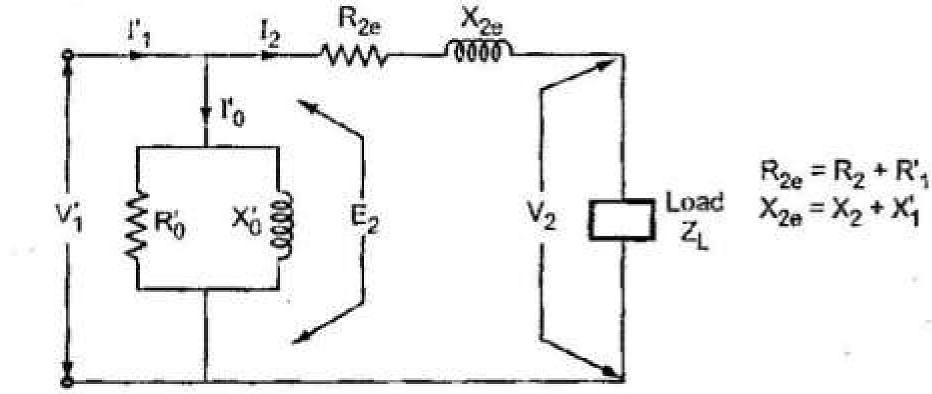


Fig:2.16Simplified equivalent circuit

<u>Approximate Voltage Drop in a Transformer</u>



$$\vec{E_2} = \overrightarrow{I_2 R_{2e}} + \overrightarrow{I_2 X_{2e}} + \vec{V_2} = \vec{V_2} + \vec{I_2}(R_{2e} + jX_{2e}), \quad \vec{E_2} = \vec{V_2} + \overrightarrow{I_2 Z_{2e}}$$

Fig. 2.17

Primary parameter is referred to secondary there are no voltage drop in primary. When there is no load, $I_2 = 0$ and we get no load terminal voltage drop in

$V_{2o} = E_2$ = no load terminal voltage

V_2 = terminal voltage on load

For Lagging P.F.

i) The current I_2 lags V_2 by angle ϕ_2

ii) Take V_2 as reference

iii) I_2R_{2e} is in phase with I_2 while $I_2 X_{2e}$ leads I_2 by 90˚

iv) Draw the circle with O as centre and OC as radius cutting extended OA at M. as OA = V_2 and now OM = E_2.

v) The total voltage drop is AM = I_2Z_{2e}.

vi) The angle α is practically very small and in practice M&N are very close to each other. Due to this the approximate voltage drop is equal to AN instead of AM

AN – approximate voltage drop

To find AN by adding AD& DN

$AD = AB \cos\phi = I_2R_{2e} \cos\phi$

$DN = BL \sin\phi = I_2X_{2e} \sin\phi$

$AN = AD + DN = I_2R_{2e} \cos\phi_2 + I_2X_{2e} \sin\phi_2$

Assuming: $\phi_2 = \phi_1 = \phi$

Approximate voltage drop = $I_2R_{2e} \cos\phi + I_2X_{2e} \sin\phi$ (referred to secondary)

Similarly: Approximate voltage drop = $I_1R_{1e} \cos\phi + I_1X_{1e} \sin\phi$ (referred to primary)

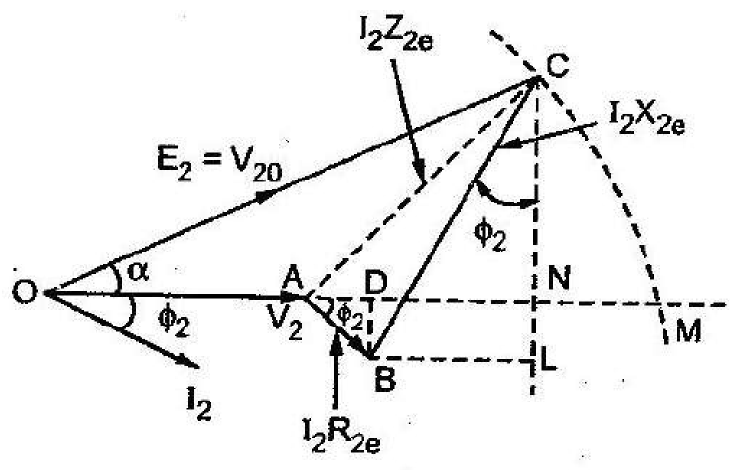


Fig:2.18

For Leading P.F Loading

I_2 leads V_2 by angle ϕ_2

Approximate voltage drop = $I_2R_{2e} \cos\phi - I_2X_{2e} \sin\phi$ (referred to secondary)

Similarly: Approximate voltage drop = $I_1R_{1e} \cos\phi - I_1X_{1e} \sin\phi$ (referred to primary)

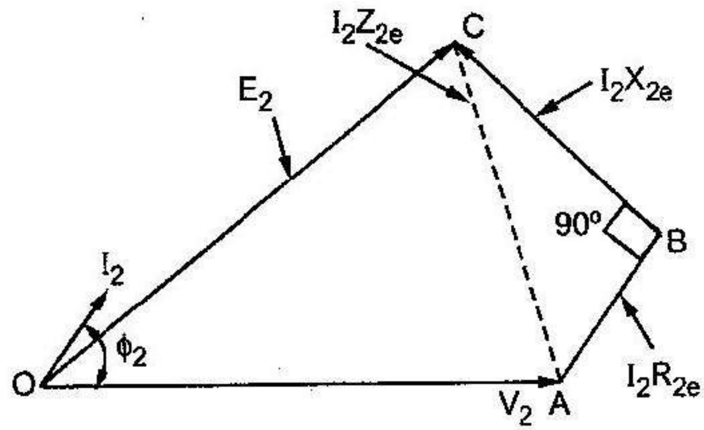
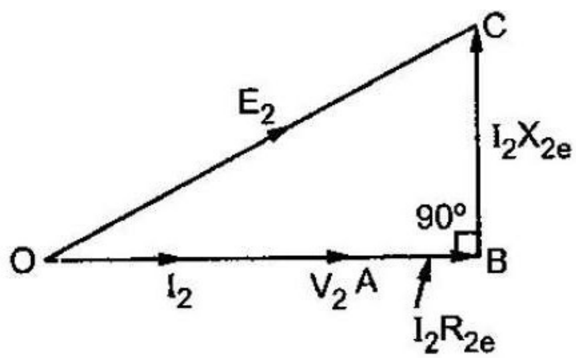


Fig: 2.19

For Unity P.F. Loading

Approximate voltage drop = I_2R_{2e} (referred to secondary)

Similarly: Approximate voltage drop = I_1R_{1e} (referred to primary)



$\cos\phi = 1$
$\sin\phi = 0$

Fig: 2.20

Approximate voltage drop = $E_2 - V_2$

$$= I_2R_{2e}\cos\phi \pm I_2X_{2e}\sin\phi \quad \text{(referred to secondary)}$$

$$= I_1R_{1e}\cos\phi \pm I_1X_{1e}\sin\phi \quad \text{(referred to primary)}$$

Losses in a Transformer

The power losses in a transformer are of two types, namely;

1. Core or Iron losses

2. Copper losses

These losses appear in the form of heat and produce (i) an increase in Temperature and (ii) a drop in efficiency.

Core or Iron losses (Pi)

These consist of hysteresis and eddy current losses and occur in the transformer core due to the alternating flux. These can be determined by open-circuit test.

Hysteresis loss = $k_h\ f\ Bm^{1.6}$ watts /m^3

K_h – hysteresis constant depend on material

f - Frequency

B_m – maximum flux density

Eddy current loss = $K_e\ f^2\ B_m^2 t^2$ watts /m^3

Ke – eddy current constant

t - Thickness of the core

Both hysteresis and eddy current losses depend upon

(i) Maximum flux density B_m in the core

(ii) Supply frequency f. Since transformers are connected to constant-frequency, constant voltage supply, both f and Bm are constant. Hence, core or iron losses are practically the same at all loads.

Iron or Core losses, Pi = Hysteresis loss + Eddy current loss = Constant losses (Pi)

The hysteresis loss can be minimized by using steel of high silicon content .Whereas eddy current loss can be reduced by using core of thin laminations.

Copper losses (P_{cu})

These losses occur in both the primary and secondary windings due to their ohmic resistance. These can be determined by short-circuit test. The copper loss depends on the magnitude of the current flowing through the windings.

$$\text{Total copper loss} = I_1^2R_1 + I_2^2R_2 = I_1^2(R_1 + R_2') = I_2^2(R_2 + R_1')$$

$$\boxed{\text{Total loss} = \text{iron loss} + \text{copper loss} = Pi + Pcu}$$

Efficiency of a Transformer

Like any other electrical machine, the efficiency of a transformer is defined as the ratio of output power (in watts or kW) to input power (watts or kW) i.e.

Power output = power input – Total losses

Power input = power output + Total losses

= power output + Pi + P_{cu}

$$\text{Efficiency} = \frac{power\ output}{power\ input}$$

$$\text{Efficiency} = \frac{power\ output}{power\ input + Pi + Pcu}$$

Power output = $V_2I_2\cos\phi$, $\cos\phi$ = load power factor

Transformer supplies full load of current I_2 and with terminal voltage V_2

Pcu = copper losses on full load = $I_2^2 R_{2e}$

$$\text{Efficiency} = \frac{V2I2\cos\phi}{V2I2\cos\phi + Pi + I2^2R2e}$$

V_2I_2 = VA rating of a transformer

$$\text{Efficiency} = \frac{(\text{VA rating})\text{X}\cos\phi}{(\text{VA rating})\text{X}\cos\phi + Pi + I2^2R2e}$$

$$\boxed{\%\ \text{Efficiency} = \frac{(\text{VA rating})\text{X}\cos\phi}{(\text{VA rating})\text{X}\cos\phi + Pi + I2^2R2e}\ \text{X}\ 100}$$

This is full load efficiency and I_2 = full load current.

We can now find the full-load efficiency of the transformer at any p.f. without actually loading the transformer.

$$\text{Full load Efficiency} = \frac{(\text{Full load VA rating})\text{X}\cos\phi}{(\text{Full load VA rating})\text{X}\cos\phi + Pi + I2^2R2e}$$

Also for any load equal to n x full-load,

Corresponding total losses = $P_i + n^2 P_{Cu}$

$$n = \text{fractional by which load is less than full load} = \frac{actual\ load}{full\ load}$$

$$n = \frac{half\ load}{fullload} = \frac{(\frac{1}{2})}{1} = 0.5$$

$$\boxed{\text{Corresponding (n) \% Efficiency} = \frac{n(\text{VA rating})\text{X}\cos\phi}{n(\text{VA rating})\text{X}\cos\phi + Pi + n^2Pcu}\ \text{X } 100}$$

Condition for Maximum Efficiency

Voltage and frequency supply to the transformer is constant the efficiency varies with the load. As load increases, the efficiency increases. At a certain load current, it loaded further the efficiency start decreases as shown in fig. 2.21.

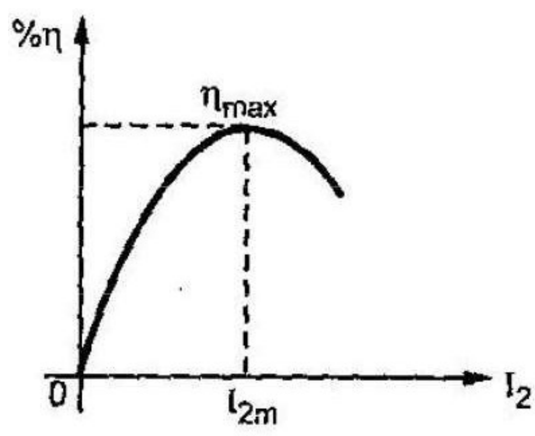


Fig: 2.21

The load current at which the efficiency attains maximum value is denoted as I_{2m} a n d maximum efficiency is denoted as ηmax, now we find -

(a) condition for maximum efficiency

(b) load current at which η_{max} occurs

(c) KVA supplied at maximum efficiency

Considering primary side,

Load output = $V_1 I_1 \cos\phi 1$

Copper loss = $I_1^2 R_{1e}$ or $I2^2 R2e$

Iron loss = hysteresis + eddy current loss = P_i

$$\text{Efficiency} = \frac{V1I1\cos\phi1 - \text{losses}}{V1I1\cos\phi1} = \frac{V1I1\cos\phi1 - I1^2R1e + Pi}{V1I1\cos\phi1}$$

$$= 1 - \frac{I1R1e}{V1I21\cos\phi1} = \frac{Pi}{V1I1\cos\phi1}$$

Differentiating both sides with respect to I2, we get

$$\frac{d\eta}{dI2} = 0 - \frac{R1e}{V1\cos\phi1} = \frac{Pi}{V1I1^2\cos\phi1}$$

For η to be maximum, $\frac{d\eta}{dI2} = 0$. Hence, the above equation becomes

$$\frac{R1e}{V1\cos\phi1} = \frac{Pi}{V1I1^2\cos\phi1} \text{ or } Pi = I_1^2R_{1e}$$

P_{cu} loss = P_i iron loss

The output current which will make P_{cu} loss equal to the iron loss. By proper design, it is possible to make the maximum efficiency occur at any desired load.

Load current I2m at maximum efficiency

For η_{max} $I_2^2 R_{2e} = P_i$ but $I_2 = I_{2m}$

$I^2_{2m} R_{2e} = Pi$ $\qquad\qquad$ $I_{2m} = \sqrt{\frac{Pi}{R2e}}$

This is the load current at η_{max}

(I_2)F.L = full load current

$$\frac{I2m}{(I2)F.L} = \frac{1}{(I2)FL}\sqrt{\frac{Pi}{R2e}}$$

$$\frac{I2m}{(I2)F.L} = \sqrt{\frac{Pi}{[(I2)F.L]^2 R2e}} = \sqrt{\frac{Pi}{[Pcu]F.L}}$$

$$\boxed{I_{2m} = (I_2)\ F.L.\ \sqrt{\frac{Pi}{[Pcu]F.L}}}$$

This is the load current at η_{max} in terms of full load current

KVA Supplied at Maximum Efficiency

For constant V_2 the KVA supplied is the function of load current.

$$\text{KVA at } \eta_{max} = I_{2m} V_2 = V_2(I_2)\text{F.L. X} \sqrt{\frac{Pi}{[Pcu]\text{F.L}}}$$

$$\boxed{\text{KVA at } \eta_{max} = (\text{KVA rating}) \text{ X} \sqrt{\frac{Pi}{[Pcu]\text{F.L}}}}$$

Substituting condition for η_{max} in the expression of efficiency, we can write expression for η_{max} as ,

as Pcu = Pi

$$\% \ \eta_{max} = \frac{V2 \ I2m \cos\phi}{V2 \ I2m\cos\phi + 2Pi} \text{ X } 100$$

$$\% \ \eta_{max} = \frac{KVA \ for \eta \max \cos\varphi}{KVA \ for \eta \max \cos\varphi + 2Pi}$$

12.8.4 All Day Efficiency (Energy Efficiency)

In electrical power system, we are interested to find out the all-day efficiency of any transformer because the load at transformer is varying in the different time duration of the day. So all day efficiency is defined as the ratio of total energy output of transformer to the total energy input in 24 hours.

$$\text{All day efficiency} = \frac{\text{kWh output during a day}}{\text{kWh input during the day}}$$

Here, kWh is kilowatt hour.

12.9 Testing of Transformer

The testing of transformer means to determine efficiency and regulation of a transformer at any load and at any power factor condition.

There are two methods

i) Direct loading test

ii) Indirect loading test

a. Open circuit test

b. Short circuit test

i) Load test on transformer

This method is also called as direct loading test on transformer because the load is directly connected to the transformer. We required various meters to measure the input and output reading while change the load from zero to full load. Fig. 2.22 shows the connection of transformer for direct load test. The primary is connected through the variac to change the input voltage as we required. Connect the meters as shown in the figure below.

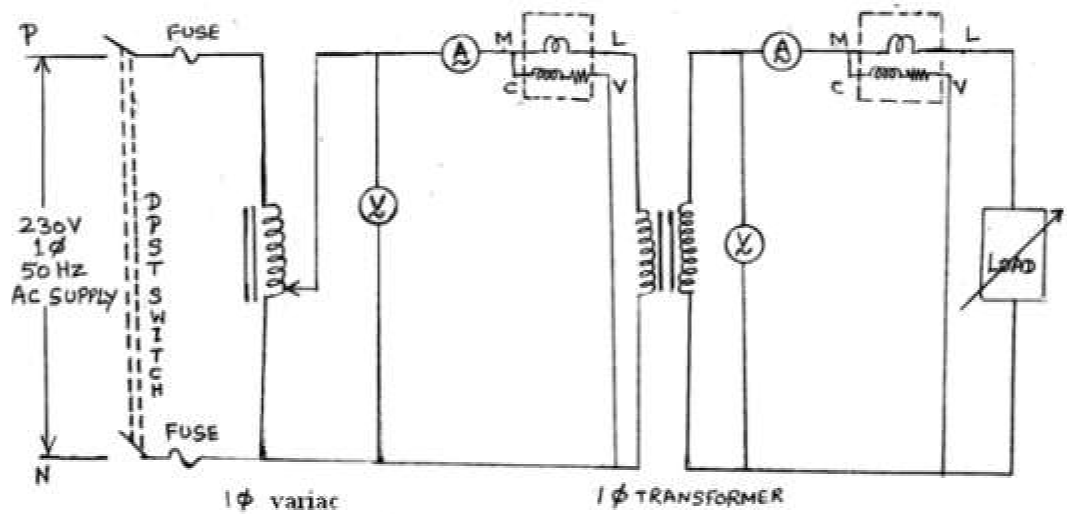


Fig: 2.22

The load is varied from no load to full load in desired steps. All the time, keep primary voltage V_1 constant at its rated value with help of variac and tabulated the reading. The first reading is to be noted on no load for which $I_2 = 0$ A and $W_2 = 0$W.

Calculation

From the observed reading

W_1 = input power to the transformer

W_2 = output power delivered to the load

$$\% \eta = \frac{W2}{W1} \text{ X } 100$$

The first reading is no load so $V2 = E_2$

The regulation can be obtained as

$$\% \text{ R} = \frac{E2 - V2}{V2} \text{ X } 100$$

The graph of % η and % R on each load against load current I_L is plotted as shown in fig. 2.23.

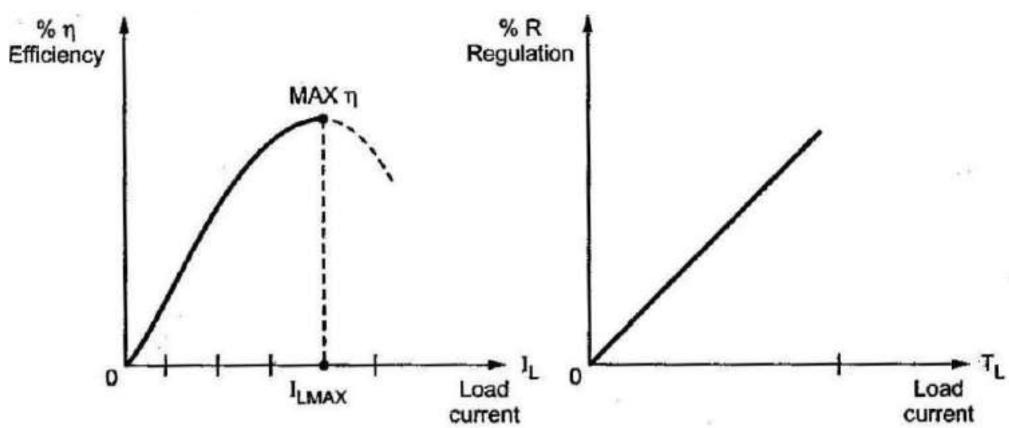


Fig: 2.23

Advantages:

1) This test enables us to determine the efficiency of the transformer accurately at any load.

2) The results are accurate as load is directly used.

Disadvantages:

1) There are large power losses during the test.

2) Load not avail in lab while test conduct for large transformer.

ii) a. Open-Circuit or No-Load Test

This test is conducted to determine the iron losses (or core losses) and parameters R_0 and X_0 of the transformer. In this test, the rated voltage is applied to the primary (usually low-voltage winding) while the secondary is left open circuited. The applied primary voltage V_1 is measured by the voltmeter, the no load current I_0 by ammeter and no-load input power W_0 by wattmeter as shown in Fig.2.24.a. As the normal rated voltage is applied to the primary, therefore, normal iron losses will occur in the transformer core. Hence wattmeter will record the iron losses and small copper loss in the primary. Since no-load current I_0 is very small (usually 2-10 % of rated current). Cu losses in the primary under no-load condition are negligible as compared with iron losses. Hence, wattmeter reading practically gives the iron losses in the transformer. It is reminded that iron losses are the same at all loads.

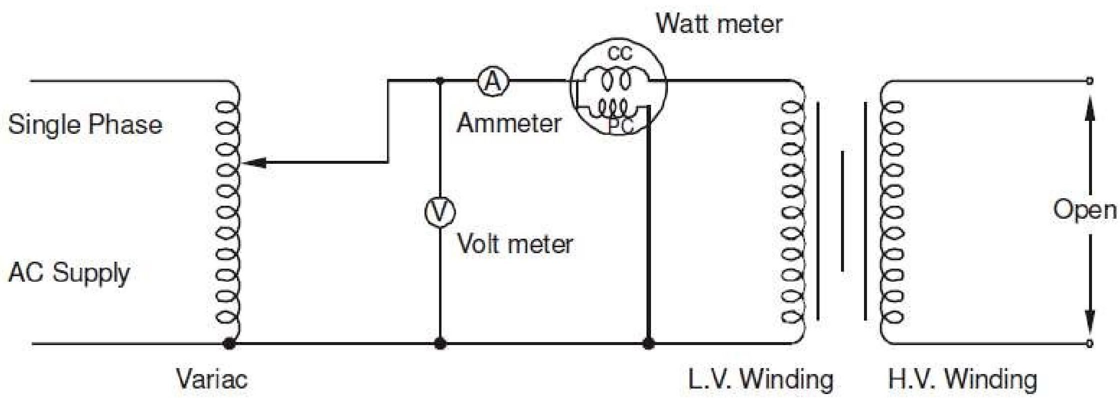


Fig: 2.24.a

Iron losses, P_i = Wattmeter reading = W_0
No load current = Ammeter reading = I_0
Applied voltage = Voltmeter reading = V_1

Input power, $W_0 = V_1 I_0 \cos \phi_0$

No - load p.f., $\cos\phi = \frac{Wo}{Vo\,Io}$ = no load power factor

$I_m = I_o \sin \phi_o$ = magnetizing component
$I_c = I_o \cos \phi_o$ = Active component

$$R_o = \frac{Vo}{Ic}\ \Omega, \quad X_o = \frac{Vo}{Im}\ \Omega$$

Under no load conditions the PF is very low (near to 0) in lagging region. By using the above data we can draw the equivalent parameter shown in Figure 2.24.b.

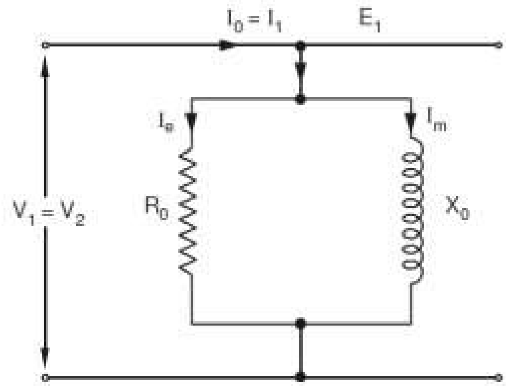


Fig: 2.24.b

Thus open-circuit test enables us to determine iron losses and parameters R_0 and X_0 of the transformer.

ii) b. Short-Circuit or Impedance Test

This test is conducted to determine R_{1e} (or R_{2e}), X_{1e} (or X_{2e}) and full-load copper losses of the transformer. In this test, the secondary (usually low-voltage winding) is short-circuited by a thick conductor and variable low voltage is applied to the primary as shown in Fig.2.25. The low input voltage is gradually raised till at voltage V_{SC}, full-load current I_1 flows in the primary. Then I_2 in the secondary also has full-load value since $I_1/I_2 = N_2/N_1$. Under such conditions, the copper loss in the windings is the same as that on full load. There is no output from the transformer under short-circuit conditions. Therefore, input power is all loss and this loss is almost entirely copper loss. It is because iron loss in the core is negligibly small since the voltage V_{SC} is very small. Hence, the wattmeter will practically register the full load copper losses in the transformer windings.

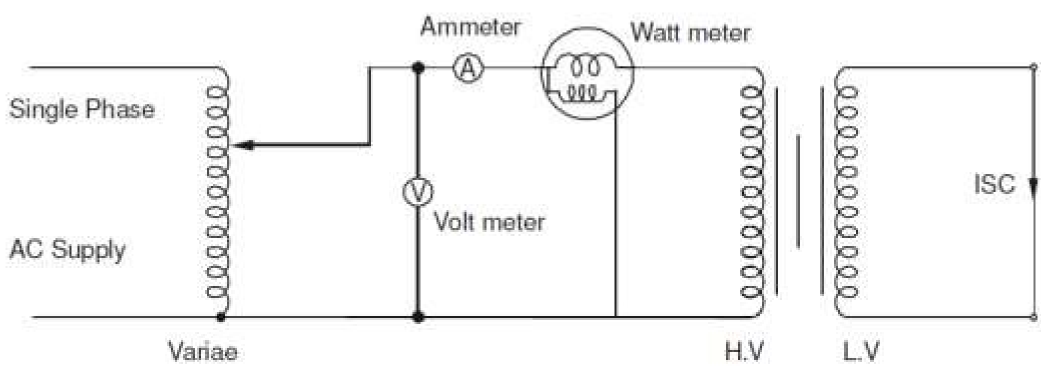


Fig: 2.25.a

Full load Cu loss, PC = Wattmeter reading = W_{sc}
Applied voltage = Voltmeter reading = V_{SC}
F.L. primary current = Ammeter reading = I_1

$Pcu = I_1^2R_1 + I_1^2 R_2' = I_1^2 R_{1e}$, $R_{1e} = \frac{Pcu}{I1^2}$
Where R_{1e} is the total resistance of transformer referred to primary.

Total impedance referred to primary, $Z1e = \sqrt{Z1e^2 - R1e^2}$,

short – circuit P.F, $\cos \Phi = \frac{Pcu}{Vsc\ I1}$ Thus short-circuit lest gives full-load Cu loss, R1e and X1e.

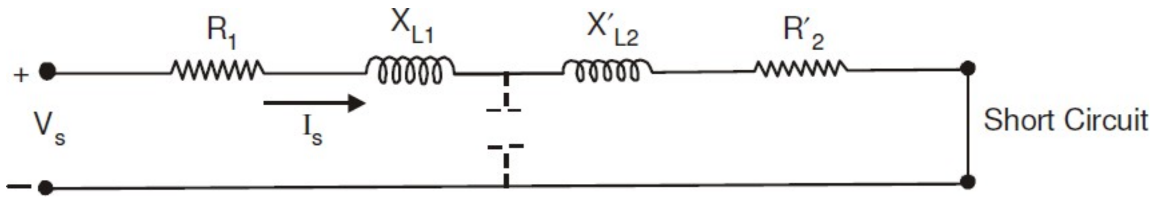


Fig: 2.25.b

From fig: 2.25.b we can calculate,

$$\text{equivalent resistance } R_{eq} = \frac{W_s}{I_s^2} = R_1 + R_2'$$

$$\text{and equivalent impedance } Z_{eq} = \frac{V_s}{I_s}$$

So we calculate equivalent reactance

$$X_{eq} = \sqrt{Z_{eq}^2 - R_{eq}^2} = X_{L1} + X'_{L2}$$

These R_{eq} and X_{eq} are equivalent resistance and reactance of both windings referred in HV side. These are know as equivalent circuit resistance and reactance.

Voltage Regulation of Transformer

Under no load conditions, the voltage at the secondary terminals is E_2 and

$$E_2 \approx V_1 \cdot \frac{N_2}{N_1}$$

(This approximation neglects the drop R_1 and X_{l1} due to small no load current). As load is applied to the transformer, the load current or the secondary current increases. Correspondingly, the primary current

I_1 also increases. Due to these currents, there is a voltage drop in the primary and secondary leakage reactances, and as a consequence the voltage across the output terminals or the load terminals changes. In quantitative terms this change in terminal voltage is called Voltage Regulation.

Voltage regulation of a transformer is defined as the drop in the magnitude of load voltage (or secondary terminal voltage) when load current changes from zero to full load value. This is expressed as a fraction of secondary rated voltage.

$$\text{Regulation} = \frac{\text{Secondary terminal voltage at no load} - \text{Secondary terminal voltage at any load}}{\text{Secondary rated voltage}}$$

The secondary rated voltage of a transformer is equal to the secondary terminal voltage at no load (i.e. E_2), this is as per IS.

Voltage regulation is generally expressed as a percentage.

$$\text{Percent voltage regulation (\% VR)} = \frac{E_2 - V_2}{E_2} \times 100\,.$$

Note that E_2, V_2 are magnitudes, and not phasor or complex quantities. Also note that voltage regulation depends not only on load current, but also on its power factor. Using approximate equivalent circuit referred to primary or secondary, we can obtain the voltage regulation. From approximate equivalent circuit referred to the secondary side and phasor diagram for the circuit.

$$E_2 = V_2 + I_2\, r_{eq} \cos\phi_2 \pm I_2\, x_{eq} \sin\phi_2$$

where $r_{eq} = r_2 + r_1^1$ (referred to secondary) $x_e = x_2 + x_1^1$ (+ sign applies lagging power factor load and – sign applies to leading pf load).

So $$\frac{E_2 - V_2}{E_2} = \frac{I_2\ r_{eq}\ \cos\phi_2 \pm I_2\ x_{eq}\ \sin\phi_2}{E_2}$$

$$\frac{E_2 - V_2}{E_2} = \frac{I_2\ r_{eq}}{E_2}\cos\phi_2 \pm \frac{I_2\ x_{eq}}{E_2}\sin\phi_2$$

% Voltage regulation = (% resistive drop) $\cos\phi_2 \pm$ (% reactive drop) $\sin\phi_2$.

Ideally voltage regulation should be zero.

Auto-transformers

The transformers we have considered so far are two-winding transformers in which the electrical circuit connected to the primary is electrically isolated from that connected to the secondary. An auto-transformer does not provide such isolation, but has economy of cost combined with increased efficiency. Fig.2.26 illustrates the auto-transformer which consists of a coil of N_A turns between terminals 1 and 2, with a third terminal 3 provided after N_B turns. If we neglect coil resistances and leakage fluxes, the flux linkages of the coil between 1 and 2 equals $N_A\ \phi m$ while the portion of coil between 3 and 2 has a flux linkage $N_B\ \phi m$. If the induced voltages are designated as E_A and E_B, just as in a two winding transformer,

$$\frac{\bar{E}_A}{\bar{E}_B} = \frac{N_A}{N_B}$$

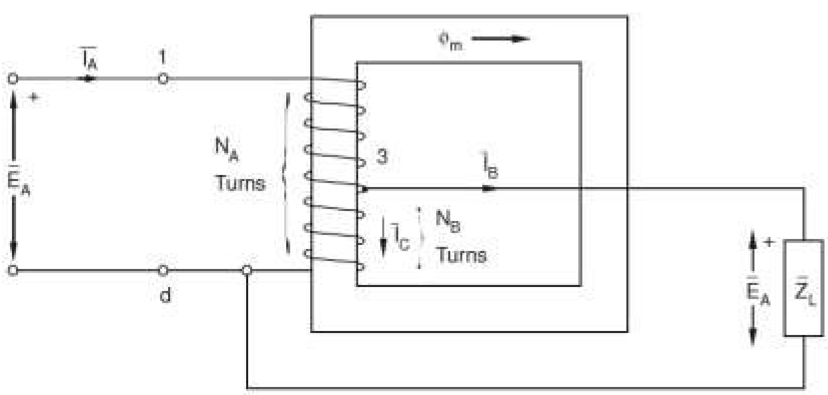


Fig: 2.26

Neglecting the magnetizing ampere-turns needed by the core for producing flux, as in an ideal transformer, the current I_A flows through only $(N_A - N_B)$ turns. If the load current is I_B, as shown by

Kirchhoff's current law, the current I_C flowing from terminal 3 to terminal 2 is (I_A - I_B). This current flows through N_B turns. So, the requirement of a net value of zero ampere-turns across the core demands that

$$(N_A - N_B)\,\bar{I}_A + (\bar{I}_A - \bar{I}_B)\,N_B = 0$$

or

$$N_A\,\bar{I}_A - N_B\,\bar{I}_B = 0$$

Hence, just as in a two-winding transformer,

$$\frac{\bar{I}_A}{\bar{I}_B} = \frac{N_B}{N_A}$$

Consequently, as far as voltage, current converting properties are concerned, the autotransformer of Figure: 26 behaves just like a two-winding transformer. However, in the autotransformer we don't need two separate coils, each designed to carry full load values of current.

Parallel Operation of Transformers

It is economical to install numbers of smaller rated transformers in parallel than installing a bigger rated electrical power transformers. This has mainly the following advantages,

To maximize electrical power system efficiency: Generally electrical power transformer gives the maximum efficiency at full load. If we run numbers of transformers in parallel, we can switch on only those transformers which will give the total demand by running nearer to its full load rating for that time. When load increases, we can switch none by one other transformer connected in parallel to fulfil the total demand. In this way we can run the system with maximum efficiency.

To maximize electrical power system availability: If numbers of transformers run in parallel, we can shut down any one of them for maintenance purpose. Other parallel transformers in system will serve the load without total interruption of power.

To maximize power system reliability: if any one of the transformers run in parallel, is tripped due to fault of other parallel transformers is the system will share the load, hence power supply may not be interrupted if the shared loads do not make other transformers over loaded.

To maximize electrical power system flexibility: There is always a chance of increasing or decreasing future demand of power system. If it is predicted that power demand will be increased in future, there must be a provision of connecting transformers in system in parallel to fulfil the extra demand because, it is not economical from business point of view to install a bigger rated single transformer by forecasting the increased future demand as it is unnecessary investment of money. Again if future demand is decreased, transformers running in parallel can be removed from system to balance the capital investment and its return.

Conditions for Parallel Operation of Transformers

When two or more transformers run in parallel, they must satisfy the following conditions for satisfactory performance. These are the conditions for parallel operation of transformers.

- ***Same voltage ratio of transformer.***
- ***Same percentage impedance.***
- ***Same polarity.***
- ***Same phase sequence.***
- ***Same Voltage Ratio***

Same voltage ratio of transformer.

If two transformers of different voltage ratio are connected in parallel with same primary supply voltage, there will be a difference in secondary voltages. Now say the secondary of these transformers are connected to same bus, there will be a circulating current between secondaries and therefore between primaries also. As the internal impedance of transformer is small, a small voltage difference may cause sufficiently high circulating current causing unnecessary extra I2R loss.

Same Percentage Impedance

The current shared by two transformers running in parallel should be proportional to their MVA ratings. Again, current carried by these transformers are inversely proportional to their internal impedance. From these two statements it can be said that, impedance of transformers running in parallel are inversely proportional to their MVA ratings. In other words, percentage impedance or per unit values of impedance should be identical for all the transformers that run in parallel.

Same Polarity

Polarity of all transformers that run in parallel, should be the same otherwise huge circulating current that flows in the transformer but no load will be fed from these transformers. Polarity of transformer means the instantaneous direction of induced emf in secondary. If the instantaneous directions of induced secondary emf in two transformers are opposite to each other when same input power is fed to both of the transformers, the transformers are said to be in opposite polarity. If the instantaneous directions of induced secondary e.m.f in two transformers are same when same input power is fed to the both of the transformers, the transformers are said to be in same polarity.

Same Phase Sequence

The phase sequence or the order in which the phases reach their maximum positive voltage, must be identical for two parallel transformers. Otherwise, during the cycle, each pair of phases will be short circuited.

The above said conditions must be strictly followed for parallel operation of transformers but totally identical percentage impedance of two different transformers is difficult to achieve practically, that is why the transformers run in parallel may not have exactly same percentage impedance but the values would be as nearer as possible.

Why Transformer Rating in kVA?

An important factor in the design and operation of electrical machines is the relation between the life of the insulation and operating temperature of the machine. Therefore, temperature rise resulting from the losses is a determining factor in the rating of a machine. We know that copper loss in a transformer depends on current and iron loss depends on voltage. Therefore, the total loss in a transformer depends on the volt-ampere product only and not on the phase angle between voltage and current i.e., it is independent of load power factor. For this reason, the rating of a transformer is in kVA and not kW.

Acknowledgement

The committee members gratefully acknowledge google, scribd, NPTEL, openoffice, sumantra pdf, scilab for myriad suggestions and help for preparing this lecture note. The committee members also wants to express their gratitude to the persons out there who thinks knowledge should be free and be accessible and sharable without any restrictions so that every single person on this planet has the same opportunity to explore, expand and become enlightened by the collective gifts of mankind.

However apart from this lecture note students/readers are strongly recommended to follow the below mentioned books in the references and above all confer with the faculty for thorough knowledge of this authoritative subject of electrical engineering.

References

[1] P.S. Bimbhra, Electrical Machinery, Khanna Pub, 7th Ed.

[2] P.S. Bimbhra, Generalized Theory of Electrical Machine, Khanna Pub., 2nd Ed.

[3] M.G. Say, The Performance and Design of Alternating Current Machines, CBS Pub., 3rd Ed.

[4] A.E.Fitzgerald, S.D.Umans, C.Kingsley Jr., Electric Machinery, McGraw Hill Education(india) pvt. Ltd., 6th Ed.

[5] A.S. langsdorf, Theory of Alternating Current Machinery, McGraw Hill Education(india) pvt. Ltd., 2nd Ed.

[6]T.C. Lloyd, A.F. Puchstein and A.G. Conrad, Alternating-Current Machines, John Wiley& Sons pub., 3rd Ed.

[7] Irving Kosow, Electric Machinery and Transformers, Pearson India, 2nd Ed.

[8] C.L.Dawes, A course in Electrical Engineering, Vol-I, Vol-II, Mcgraw Hill Book Company Inc.

[9] P.K. Mukherjee & S. Chakravorty, Electrical Machines, Dhanpat Rai Publications, 2nd Ed.

[10] I.J Nagrath & D.P.Kothari, Electrical Machines, McGraw Hill Education(india) pvt. Ltd., 4th Ed.

[11] N.N.Parker Smith, Problems in Electrical Engineering, CBS Pub. 9th Ed.

<u>Best of Luck to All the Students</u>

Principle of Operation

DC motor operates on the principle that when a current carrying conductor is placed in a magnetic field, it experiences a mechanical force given by F = BIL newton. Where 'B' = flux density in wb, 'I' is the current and 'L' is the length of the conductor. The direction of force can be found by Fleming's left hand rule. From the point of construction, there is no difference between a DC generator and DC motor. Figure 3.1 shows a multipolar DC motor. Armature conductors are carrying current downwards under North Pole and upwards under South Pole. When the field coils are excited, with current carrying armature conductors, a force is experienced by each armature conductor whose direction can be found by Fleming's left hand rule. This is shown by arrows on top of the conductors. The collective force produces a driving torque which sets the armature into rotation. The function of a commutator in DC motor is to provide a continuous and unidirectional torque.

In DC generator the work done in overcoming the magnetic drag is converted into electrical energy. Conversion of energy from electrical form to mechanical form by a DC motor takes place by the work done in overcoming the opposition which is called the 'back emf'.

Back EMF

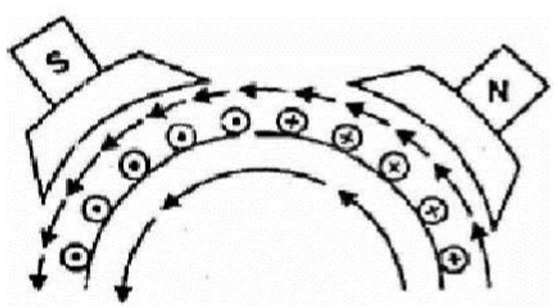


Fig. 3.1 Generation of force in DC motor

It is the dynamically induced emf in the armature conductors when the armature rotates following principle of DC motor. The direction of this induced emf can be determined using Fleming's right hand rule. This emf act in opposition to the supply voltage of the armature. It opposes the supply voltage that is why it is called back emf. The value of this induced emf is same as the value of the emf induced in dc generator. The work done in overcoming this opposition is converted into mechanical energy.

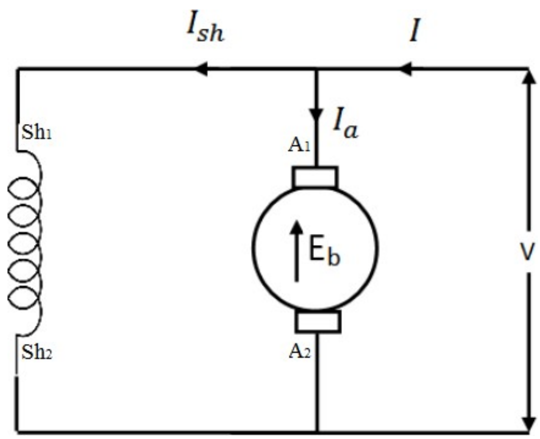


Fig. 3.2 Schematic diagram of DC shunt motor

Fig shown 3.2 a DC shunt motor the rotating armature generating the back emf E_b. The armature current can be written as

$$I_a = \frac{V_T - E_b}{r_a}$$

Where r_a is armature resistance,

$$\boxed{E_b = \frac{P\phi ZN}{60A}}$$

Armature current is proportional to back emf. So back emf is a controlling factor of armature current.

<u>Torque Equation</u>

Let T_a = armature torque in N –m developed by the armature of a motor running at N.rps.

Therefore $P = T_a \text{ X } \frac{2\pi N}{60}$ Watts

Electrical equivalent of mechanical power developed $P_m = E_b I_a$

$$P_m = E_b I_a = P = T_a \times \frac{2\pi N}{60}$$

$$T_a = \frac{E_b I_a}{2\pi N} \times 60 \text{ Also, on substituting for Eb i.e., } E_b = \frac{P\phi ZN}{60A}$$

Therefore,

$$T_a = \frac{I_a \times 60}{2\pi N} \times \frac{P\phi ZN}{60A}$$

$$T_a = \frac{I_a}{2\pi} \times \frac{P\phi Z}{A}$$

$$\boxed{T_a = \frac{P\phi Z I_a}{2\pi A}} \text{ N-m}$$

From the above equation for torque, it is seen that

(i) $T_a = k\phi I_a$

(ii) $T_a \propto I_a^2$ - For series motor (because $\phi \propto I_a$) before saturation. After saturation $T_a \propto I_a$

(iii) $T_a \propto I_a$ - For shunt motor. (because ϕ is constant in a shunt motor)

3.4 Characteristics of DC Motors

There are three important characteristics-

1. Armature torque vs armature current T_a vs I_a *(Electrical characteristics)*

2. Speed vs armature current characteristic N vs I_a

3. Speed vs torque N vs T_a (Mechanical characteristics)

Characteristics of DC shunt motor

Armature torque vs armature current T_a vs I_a characteristics

For a shunt motor flux can be assumed practically constant (at heavy loads, decreases, due to increased armature reaction)

$T_a = k\phi I_a$

ϕ is constant, $T_a \propto I_a$

Therefore electrical characteristic of a shunt motor is a straight line through origin shown by dotted line in figure 3.3. Armature reaction weakens the flux hence T_a vs I_a characteristic bends as shown by dark line in figure 3.3, Shunt motors should never be started on heavy loads, since it draws heavy current under such condition.

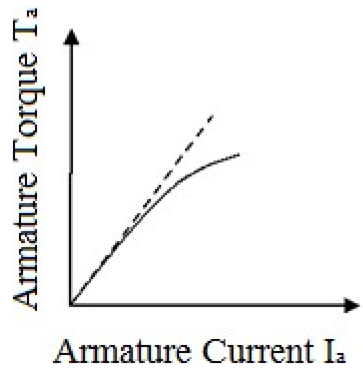


Fig. 3.3 Torque Current Characteristic of DC shunt motor

Speed vs armature current N_a vs I_a characteristics

$$N \propto \frac{E_b}{\phi}$$

$$N = \frac{V - I_a r_a}{k\phi}$$

$$N = \frac{V}{k\phi} - \frac{I_a r_a}{k\phi}$$

V is constant and in dc shunt motor ϕ is also constant. Thus with armature current speed drops and the speed current characteristics is drooping in nature is shown in figure 3.4.

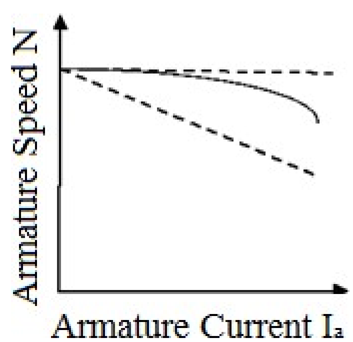


Fig. 3.4 Speed vs armature current characteristics of DC shunt motor

Speed vs armature torque N_a vs T_a characteristics

$$N = \frac{V}{k\phi} - \frac{I_a r_a}{k\phi}$$

$$I_a = \frac{T_a}{k\phi}$$

$$N = \frac{V}{k\phi} - \frac{r_a}{k\phi^2} T_a$$

Thus with increase with torque the speed of DC shunt motor decreases. The nature of the characteristics is drooping in nature shown in figure 3.5.

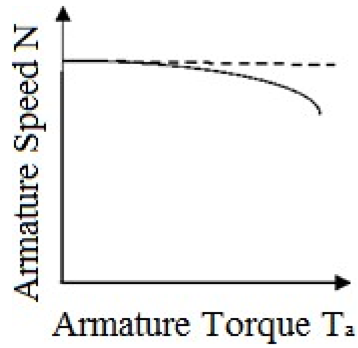


Fig. 3.5 Speed vs armature torque characteristics of DC shunt motor

Characteristics of DC series motor

Armature torque vs armature current T_a vs I_a characteristics

$T_a = k\phi I_a$

$T_a \propto I_a^2$ - For series motor (because $\phi \propto I_a$) before saturation

After saturation ϕ becomes constant thus $T_a \propto I_a$

At light loads, I_a and hence ϕ is small. But as I_a increases T_a increases as the square of the current up-to saturation. After saturation ϕ becomes constant, the characteristic becomes a straight line as shown in Figure 3.6. Therefore a series motor develops a torque proportional to the square of the armature current. This characteristic is suited where huge starting torque is required for accelerating heavy masses.

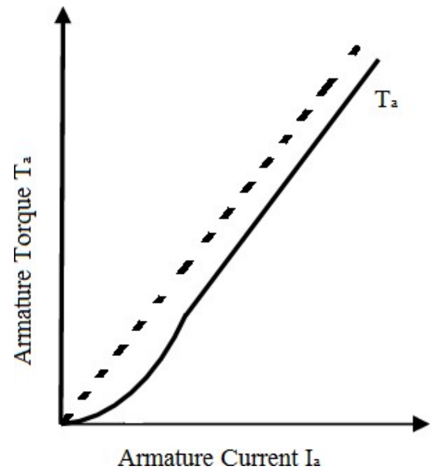


Fig. 3.6 Torque Current Characteristic of DC series motor

Speed vs armature current N_a vs I_a characteristics

$$N \propto \frac{E_b}{\phi}$$

$$N = \frac{V - I_a r_a}{k\phi}$$

$$N = \frac{V}{k\phi} - \frac{I_a r_a}{k\phi}$$

$$I_a \propto \phi$$

$$N = \frac{V}{k k_1 I_a} - \frac{k_1 r_a}{k}$$

$$N \propto \frac{1}{I_a}$$

If I_a increases, speed decreases. This characteristic is shown in figure 3.7. Therefore the speed is inversely proportional to armature current I_a. When load is heavy I_a ia heavy thus speed is low. When load is low I_a is low thus speed becomes dangerously high. Hence series motor should never started without load on it.

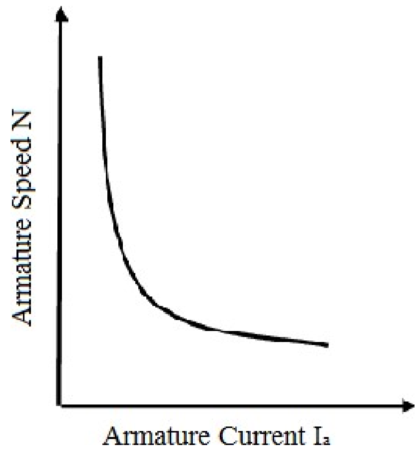


Fig. 3.7 Speed vs armature current characteristics of DC series motor

3.4.2.2 Speed vs armature torque N_a vs T_a characteristics

$N = \frac{V}{k\phi} - \frac{I_a r_a}{k\phi}$ and $\phi = k_1 I_a$

$\therefore N = \frac{V}{kk_1 I_a} - \frac{I_a r_a}{kk_1 I_a}$

$\Rightarrow N = \frac{V}{kk_1 I_a} - \frac{r_a}{kk_1}$

Now, $T_a = kI_a^2 \quad \therefore I_a = \sqrt{\frac{T_a}{k}}$

Substituting I_a

$N = \frac{V\sqrt{k}}{kk_1\sqrt{T_a}} - \frac{r_a}{kk_1}$

$\Rightarrow N = \frac{\text{Const.}}{\sqrt{T_a}} - \text{Const.}$

Thus Speed is inversely proportional to torque. The characteristics is shown in figure 3.8.

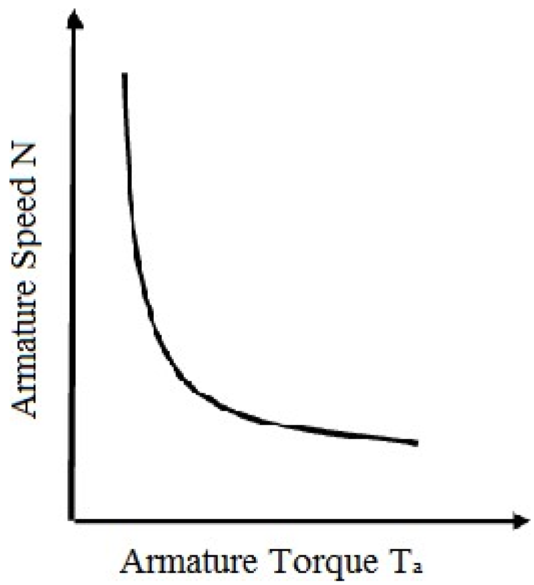


Fig. 3.8 Speed vs armature torque characteristics of DC series motor

Characteristics of DC compound motor

There are two different types of compound motors in common use, they are the cumulative compound motor and the differential compound motor. In the cumulative compound motor, the field produced by the series winding aids the field produced by the shunt winding. The speed of this motor falls more rapidly with increasing current than does that of the shunt motor because the field increases. In the differential compound motor, the flux from the series winding opposes the flux from the shunt winding. The field flux, therefore, decreases with increasing load current. Because the flux decreases, the speed may increases with increasing load. Depending on the ratio of the series-to-shunt field ampere-turns, the motor speed may increases very rapidly.

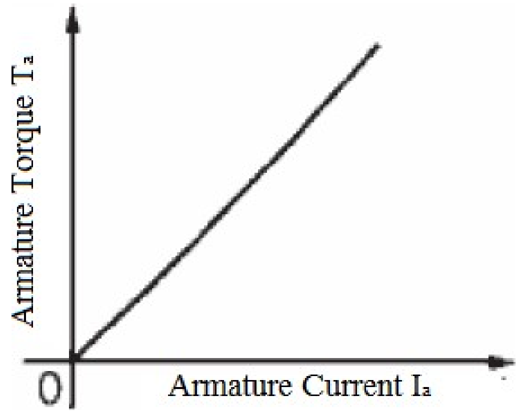


Fig. 3.9 Armature torque vs armature current characteristics of DC compound motor

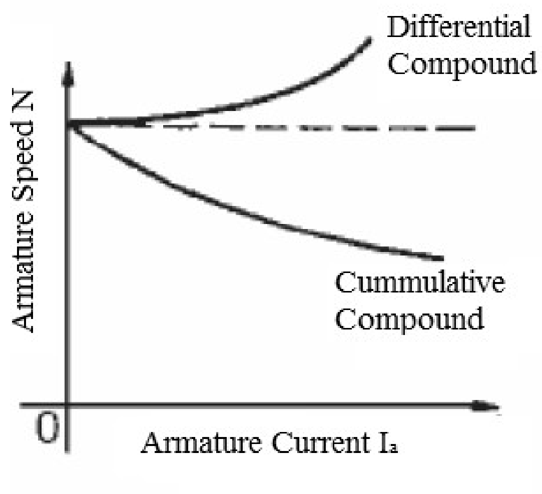


Fig. 3.10 Speed vs armature current characteristics of DC compound motors

The torque-speed (c/s) of a cumulatively compound D.C motor

In the cumulative compounded D.C. motor, there is a component of flux which is constant and another component which is proportional to its armature current (and thus to its load). Therefore, the cumulatively compounded motor has a higher starting torque than a shunt motor (whose flux is constant) but a lower starting torque than a series motor (whose entire flux is proportional to armature current). At light loads, the series field has a very small effect, so the motor behaves approximately as a shunt D.C. motor. As the load gets very large, the series flux becomes quite important and the torque-speed curve begins to look like a series motor's (c/s). A comparison of the torque-speed (c/s) of each of these type of machines is shown in figure 3.11.

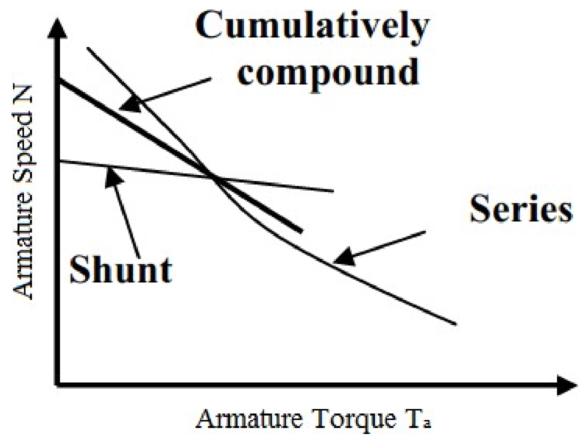


Fig. 3.11 Speed vs armature torque characteristics of DC motors

The torque-speed (c/s) of a differentially compound D.C motor

In a differentially compound D.C. motor, the shunt magneto motive force and series magneto motive force subtract from each other. This means that as the load on the motor increases, I_a increases and the flux in the motor decreases. But as the flux decreases, the speed of the motor increases. This speed increases causes anther increases in load, which further increases I_a, further decreasing the flux, and increasing the speed again. The result is that a differentially compounded motor is unstable and tends to run away. It is so bad that a differentially compounded motor is unsuitable for any application. The torque speed characteristics is shown in figure 3.12.

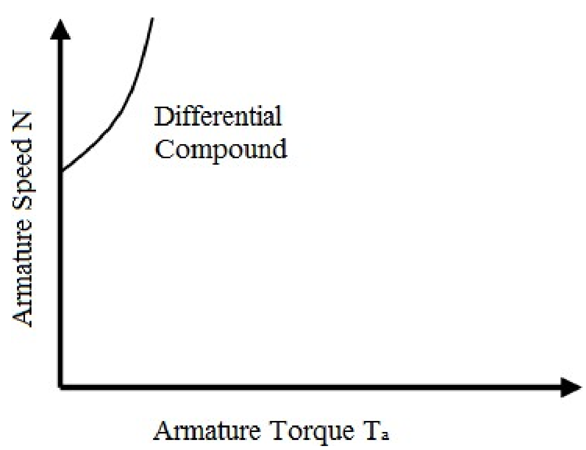


Fig. 3.12 Speed vs armature torque characteristics of DC differential compound motor

Application of DC motors

Application of DC shunt motor

The characteristics of a DC shunt motor give it a very good speed regulation, and it is classified as a constant speed motor, even though the speed does slightly decrease as load is increased. Shunt wound motors are used in industrial and automotive applications where precise control of speed and torque are required.

Application of DC series motor

For a given input current, the starting torque developed by a DC series motor is greater than that developed by a shunt motor. Hence series motors are used where huge starting torques are necessary. Ex. Cranes, hoists, electric traction etc. The DC series motor responds by decreasing its speed for the increased in load. The current drawn by the DC series motor for the given increase in load is lesser than DC shunt motor. The drop in speed with increased load is much more prominent in series motor than that in a shunt motor. Hence series motor is not suitable for applications requiring a constant speed.

Application of DC compound motor

Cumulative compound wound motors are virtually suitable for almost all applications like business machines, machine tools, agitators and mixers etc. Compound motors are used to drive loads such as shears, presses and reciprocating machines.

Differential compound motors are seldom used in practice (because of rising speed characteristics).

Armature Reaction

The action of magnetic field set up by armature current on the distribution of flux under main poles of a DC machine is called armature reaction.

When the armature of a DC machines carries current, the distributed armature winding produces its own mmf. The machine air gap is now acted upon by the resultant mmf distribution caused by the interaction of field ampere turns (AT_f) and armature ampere turns (AT_a). As a result the air gap flux density gets distorted.

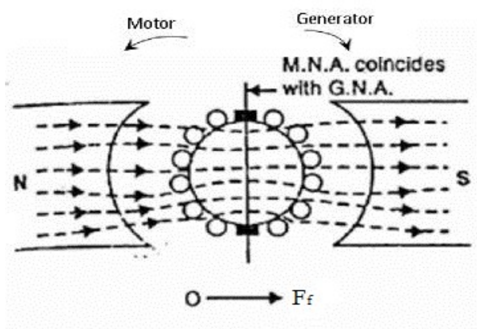


Fig. a

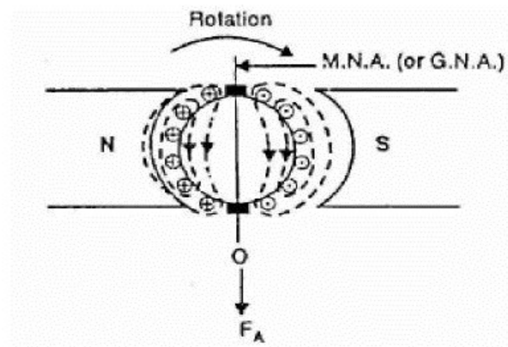


Fig. b

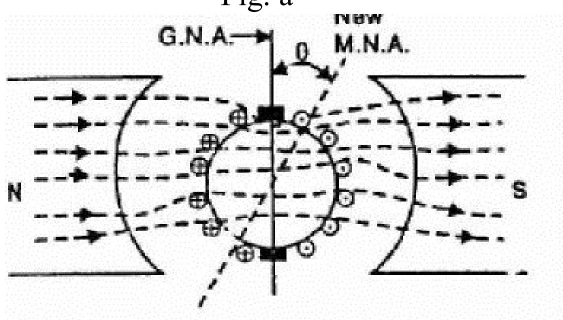


Fig. c

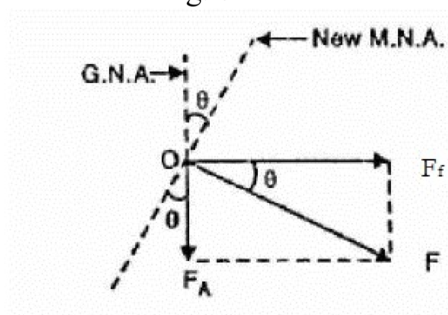


Fig. d

Figure a shows a two pole machine with single equivalent conductor in each slot and the main field mmf (F_f) acting alone. The axis of the main poles is called the direct axis (d-axis) and the interpolar axis is called quadrature axis (q-axis). It can be seen from the Figure b that armature mmf (F_a) is along the interpolar axis. F_a which is at 90^0 to the main field axis is known as cross magnetizing mmf.

Figure c shows the practical condition in which a DC machine operates when both the Field flux and armature flux are existing. Because of both fluxes are acting simultaneously, there is a shift in brush axis and crowding of flux lines at the trailing pole tip and flux lines are weakened or thinned at the leading pole tip. (The pole tip which is first met in the direction of rotation by the armature conductor is leading pole tip and the other is trailing pole tip).

If the iron in the magnetic circuit is assumed unsaturated, the net flux/pole remains unaffected by the armature reaction though the air gap flux density distribution gets distorted. If the main pole excitation is such that the iron is in the saturated region of magnetization (practical case) the increase in flux density at one end of the poles caused by armature reaction is less than the decrease at the other end, so that there is a net reduction in the flux/pole. This is called the demagnetizing effect.

Thus it can be summarized that the nature of armature reaction in a DC machine is

1. Cross magnetizing with its axis along the q-axis.
2. It causes no change in flux/pole if the iron is unsaturated but causes reduction in flux/pole in the presence of iron saturation. This is termed as demagnetizing effect. The resultant mmf 'F' is shown in figure d.

Commutation

The process of reversal of current in the short circuited armature coil is called 'Commutation'. This process of reversal takes place when coil is passing through the interpolar axis (q-axis), the coil is short circuited through commutator segments and brush.

The process of commutation of coil 'CD' is shown Fig. 3.13. In sub figure 'c' coil 'CD' carries 20A current from left to right and is about to be short circuited in figure 'd' brush has moved by a small width and the brush current supplied by the coil are as shown. In figure 'e' coil 'CD' carries no current as the brush is at the middle of the short circuit period and the brush current in supplied by coil 'AB' and coil 'EF'. In sub figure 'f' the coil 'CD' which was carrying current from left to right carries current from right to left. In sub fig 'g' spark is shown which is due to the reactance voltage. As the coil is embedded in the armature slots, which has high permeability, the coil possess appreciable amount of self inductance. The current is changed from +20 to –20. So due to self inductance and variation in the current from +20 to –20, a voltage is induced in the coil which is given by L dI/dt. This emf opposes the change in current in coil 'CD' thus sparking occurs.

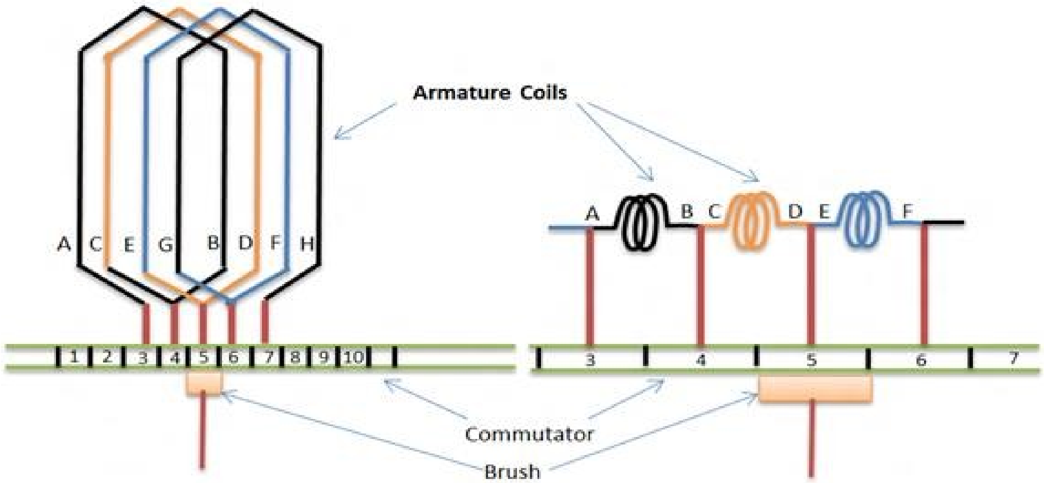


Fig a

Fig b

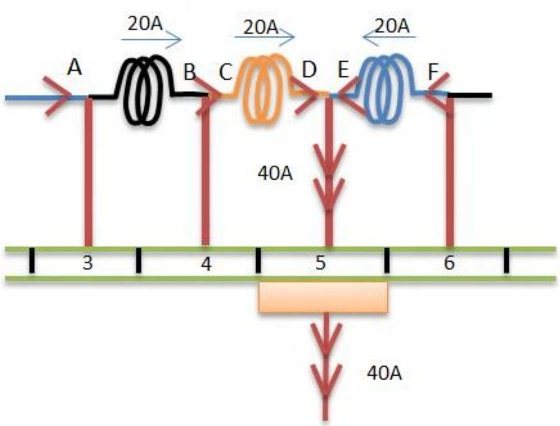


Fig.c

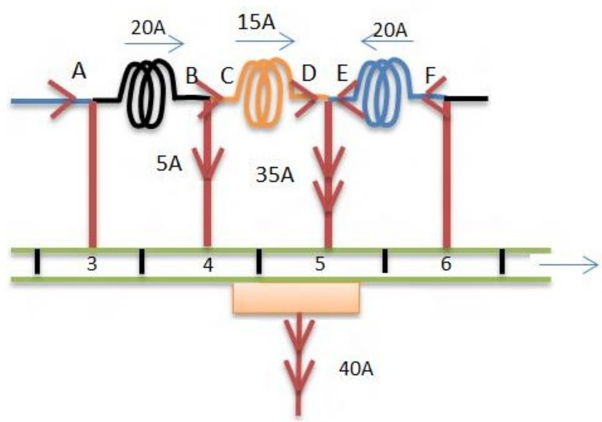


Fig.d

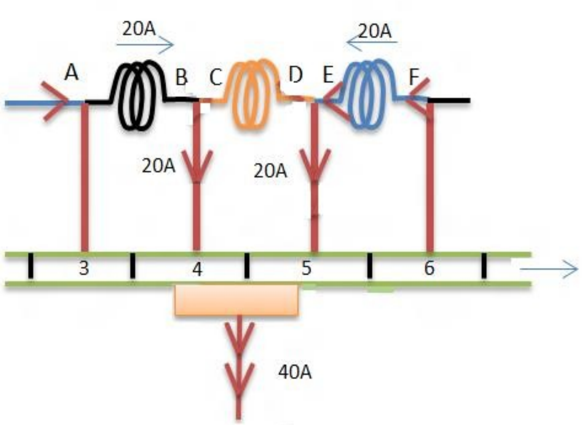


Fig.e

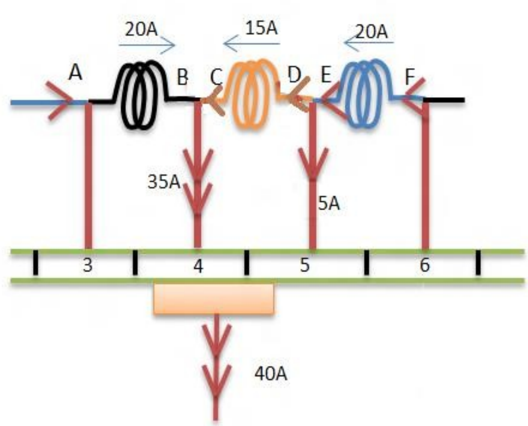


Fig.f

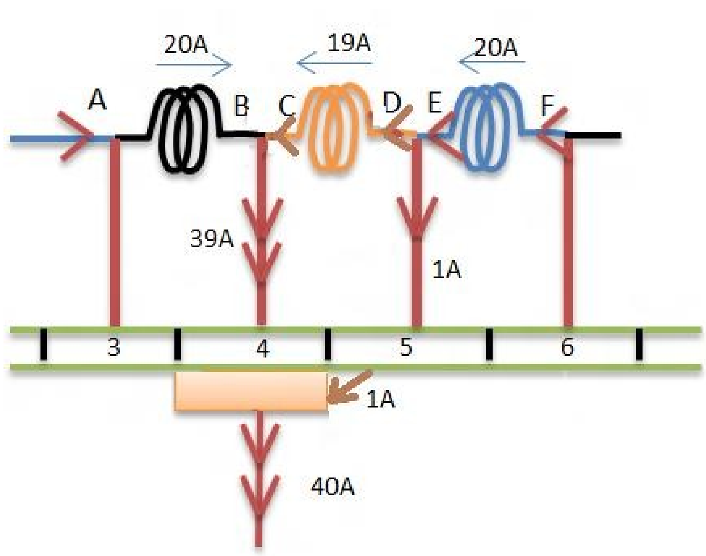


Fig.g

Fig. 3.13 a-g shows the process of commutation

Starting of DC Motor

Necessity of starter:

The current drawn by the armature is given by

$$I_a = \frac{V_T - E_b}{r_a}$$

At starting, as N=0 so Eb = 0 thus

$$I_a = \frac{V_T}{r_a}$$

Armature resistance will be very low. Therefore, the current drawn by the motor will be very high. In order to limit this high current, a starting resistance is connected in series with the armature. The starting resistance will be excluded from the circuit after the motor attains its rated speed. From there on back emf limits the current drawn by the motor.

Three Point Starter

The arrangement is shown in the figure 3.14 shows a three point starter for shunt motor.

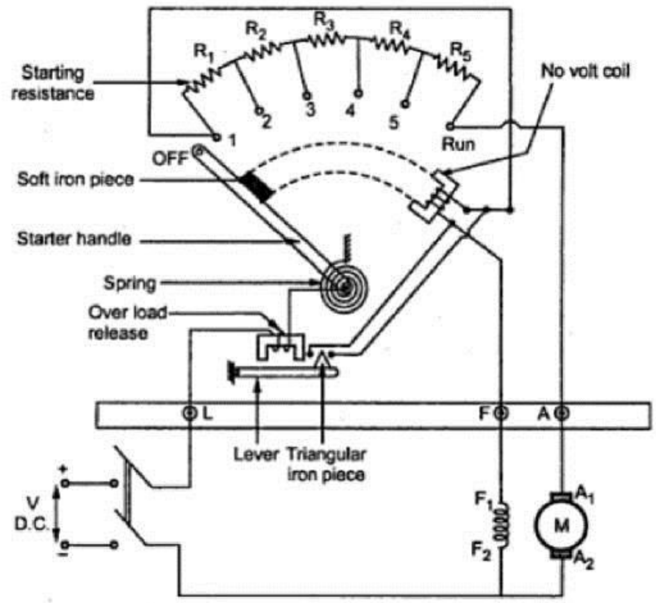


Fig. 3.14 Internal view of three point starter

It consists of resistances arranged in steps, R_1 to R_5 connected in series with the armature of the shunt motor. Field winding is connected across the supply through a protective device called 'NO – Volt Coil'. Another protection given to the motor in this starter is 'over load release coil'. To start the motor the starter handle is moved from OFF position to Run position gradually against the tension of a hinged spring. An iron piece is attached to the starter handle which is kept hold by the No-volt coil at Run position. The function of No volt coil is to get de-energized and release the handle when there is failure or disconnection or a break in the field circuit so that on restoration of supply, armature of the motor will not be connected across the lines without starter resistance. If the motor is over loaded beyond a certain predetermined value, then the electromagnet of overload release will exert a force enough to attract the lever which short circuits the electromagnet of No volt coil. Short circuiting of No volt coil results in de-energisation of it and hence the starter handle will be released and return to its off position due to the tension of the spring.

Four Point Starter

One important change is the No Volt Coil has been taken out of the shunt field and has been connected directly across the line through a Protecting resistance ‘R’. When the arm touches stud one. The current divides into three paths, 1. Through the starter resistance and the armature, 2. Through shunt field and the field rheostat and 3.Through No-volt Coil and the protecting resistance ‘R’. With this arrangement, any change of current in shunt field circuit does not affect the current passing though the NO-volt coil because, the two circuits are independent of each other. Thus the starter handle will not be released to its off position due to changes in the field current which may happen when the field resistance is varied. Fig 3.15 shows internal view of 4-point starter.

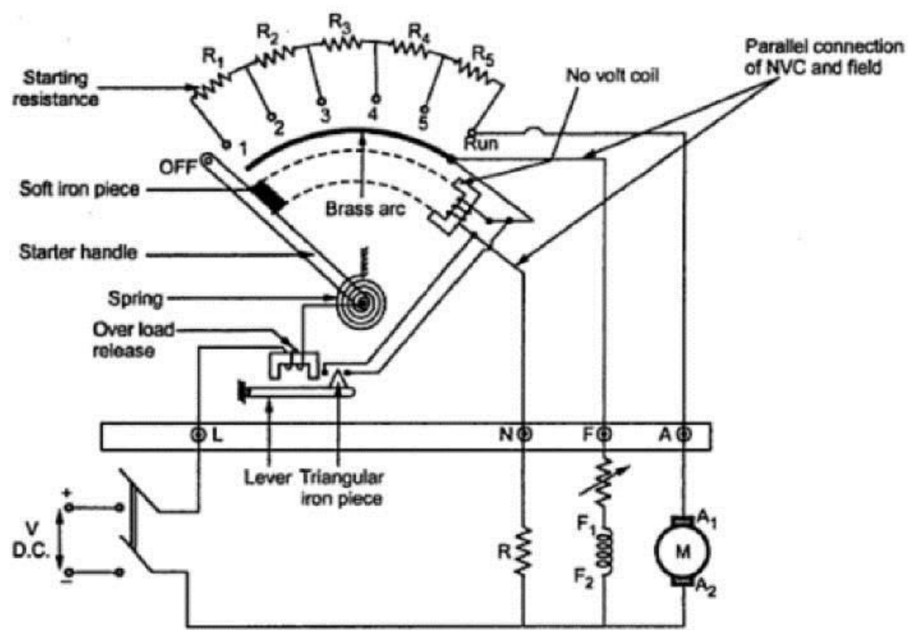


Fig. 3.15 Internal view of three point starter

DRUM CONTROLLERS

Drum controllers are used when an operator is controlling the motor directly. The drum controller is used to start, stop, reverse, and vary the speed of a motor. This type of controller is used on crane motors, elevators, machine tools, and other applications in heavy industry. As a result, the drum controller must be more rugged than the starting rheostat.

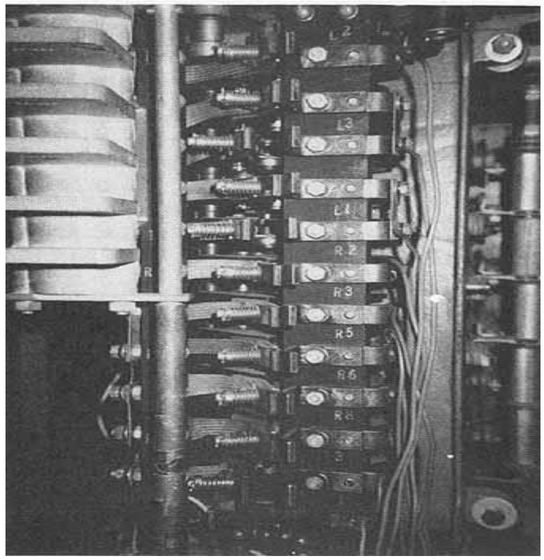


Fig. 3.16 Drum type controller shows contact fingers.

A drum controller with its cover removed is illustrated in 3.16. The switch consists of a series of contacts mounted on a movable cylinder. The contacts, which are insulated from the cylinder and from one another, are called movable contacts. There is another set of contacts, called stationary contacts, located inside the controller. These stationary contacts are arranged to touch the movable contacts as the cylinder is rotated. A handle, keyed to the shaft for the movable cylinder and contacts, is located on top of the drum controller. This handle can be moved either clockwise or counterclockwise to give a range of speed control in either direction of rotation. The handle can remain stationary in either the forward or reverse direction due to a roller and a notched wheel. A spring forces the roller into one of the notches at each successive position of the controller handle to keep the cylinder and movable contacts stationary until the handle is moved by the operator.

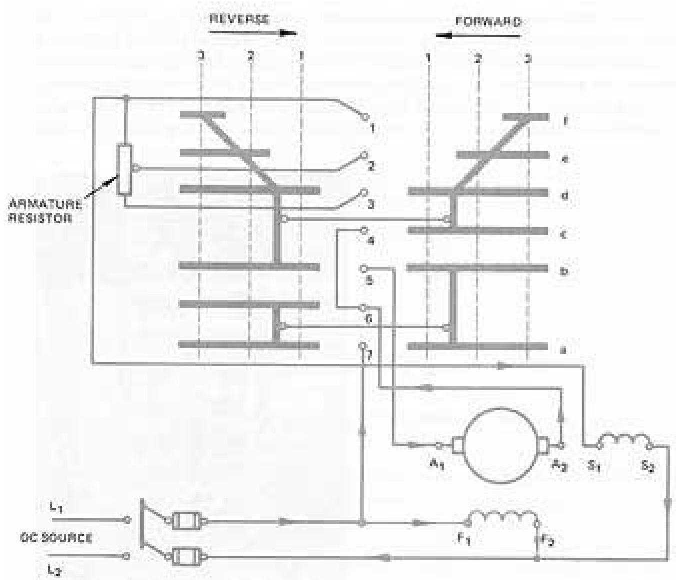


Fig. 3.17 Schematic diagram of a drum controller connected to a compound-wound motor

A drum controller with two steps of resistance is illustrated in 3.17. The contacts are represented in a flat position in this schematic diagram to make it easier to trace the circuit connections. To operate the motor in the forward direction, the set of contacts on the right must make contact with the center stationary contacts. Operation in the reverse direction requires that the set of movable contacts on the left makes contact with the center stationary contacts.

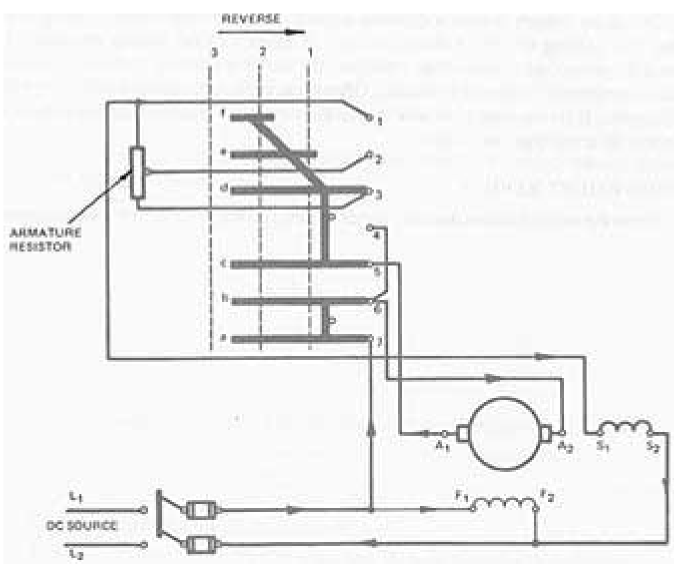


Fig. 3.18 First position of controller for reverse direction

Note in figure 3.17 that there are three forward positions and three reverse positions to which the controller handle can be set. In the first forward position, all of the resistance is in series with the armature. The circuit path for the first forward position is as follows:

1. Movable fingers a, b, c, and d contact the stationary contacts 7, 5, 4, and 3.

2. The current path is from the positive side of the line to contact 7, from 7 to a, from a to b, from b to 5, and then to armature terminal A1.

3. After passing through the armature winding to terminal A the current path is to stationary contact 6, and then to stationary contact 4.

4. From contact 4 the current path is to contact c, to d, and then to contact 3.

5. The current path then goes through the armature resistor, to the series field, and then back to the negative side of the line.

The shunt field of the compound motor is connected across the source voltage. On the second forward position of the controller handle, part of the resistance is cut out. The third forward position cuts out all of the resistance and puts the armature circuit directly across the source voltage.
In the first reverse position, all of the resistance is inserted in series with the armature. Fig. 3.18 shows the first position of the controller in the reverse direction. The current in the armature circuit's reversed. However, the current direction in the shunt and series fields is the same as the direction for the forward positions. A change in current direction in the armature only resulted in a change in the direction of rotation.

The second reverse position cuts out part of the resistance circuit. The third reverse position cuts out all of the resistance and puts the armature circuit directly across the source. Drum controllers with more positions for a greater control of speed can be obtained. However, these controllers all use the same type of circuit arrangement shown in this unit.

DC series motors require a different starting controller than shunt or compound motor. The holding circuit for the controller is in series with the starting resistance. If there is a low-voltage or no-voltage condition, the starter is returned to the off position. Drum controllers are still used frequently. Often drum controllers are used with ac as well as dc motors. It is important to be able to read the connection diagrams and the sequence diagrams on drum-type controllers.

Losses and efficiency of DC Machines

It is convenient to determine the efficiency of a rotating machine by determining the losses than by direct loading. Further it is not possible to arrange actual load for large and medium sized machines. By knowing the losses, the machine efficiency can be found by

$$\eta = \frac{\text{Output}}{\text{Output+Losses}}(\text{for Generator})$$

$$\eta = \frac{\text{Input-Losses}}{\text{Input}}(\text{for Motor})$$

In the process of energy conversion in rotating machines-current, flux and rotation are involved which cause losses in conductors, ferromagnetic materials and mechanical losses respectively.

Various losses occurring in a DC machine are listed below-

Total losses can be broadly divided into two types.

1) Constant losses

2) Variable losses

These losses can be further divided as

1) Constant losses –

i) Core loss or iron loss

a) Hysteresis loss

b) Eddy current loss

ii) Mechanical loss

a) Windage loss

b) Friction loss – brush friction loss and Bearing friction loss.

2) Variable losses –

i) copper loss ($I^2 r$)

a) Armature copper loss
b) Field copper loss
c) Brush contact loss

ii) Stray load loss

a) Copper stray load loss
b) Core stray load loss

Core loss or iron loss occurs in the armature core is due to the rotation of armature core in the magnetic flux produced by the field system. Iron loss consists of a) Hysteresis loss and b) Eddy current loss.

Hysteresis loss: This loss is due to the reversal of magnetization of armature core as the core passes under north and south poles alternatively. This loss depends on the volume and grade of iron, maximum value of flux density and frequency. Hysteresis loss is given by Steinmetz formula.

$W_h = K_h B_m^{1.6} f V$ Joule/sec or watt

Where K_h = Constant of proportionality- depends on core material.

B_m = Maximum flux density in Wb/m^2

f = Frequency in Hz

V = Volume of the armature core in m^3

Eddy Current Loss: Eddy currents are the currents set up by the induced emf in the armature core when the core cuts the magnetic flux. The loss occurring due to the flow of eddy current is known as eddy current loss. To reduce this loss the core is laminated, stacked and riveted. These laminations are insulated from each other by a thin coating of varnish. The effect of lamination is to reduce the current path because of increased resistance due to reduced cross section area of laminated core. Thus the magnitude of eddy current is reduced resulting in the reduction of eddy current loss.

Eddy Current loss is given by

$W_e = K_e B_m^2 f^2 t^2 V$ Watt

Where K_e = Constant of proportionality

B_m = Maximum flux density in Wb/m^2

f = Frequency in Hz

V = Volume of the armature core in m^3

t = Thickness of the lamination in meters

ii) Mechanical loss: these losses include losses due to windage, brush friction and bearing friction losses.

2) **Variable losses**: Variable losses consist of

(i) Copper loss:

a) Armature copper loss: This loss occurs in the armature windings because of the resistance of armature windings, when the current flows through them. The loss occurring is termed as copper loss or r loss. This loss varies with the varying load.

b) Field copper loss: This is the loss due to current flowing in the field windings of the machine. c) Brush contact drop: This is due the contact resistance between the brush and the commutator. This loss remains constant with load.

(ii) Stray load loss: The additional losses which vary with the load but cannot be related to current in a simple manner are called stray load loss. Stray load losses are.

Copper stray load loss: the loss occurring in the conductor due to skin effect and loss due to the eddy currents in the conductor set up by the flux passing through them are called copper stray load loss.

Core stray load loss: When the load current flows through the armature conductors, the flux density distribution gets distorted in the teeth and core. The flux density decreases at one end of the flux density wave and increases at the other. Since the core loss is proportional to the square of the flux density, the decrease in flux density will be less than the increase due to the increase in flux density, resulting in a net increase in the core loss predominantly in the teeth, is known as stray load loss in the core.

Further under highly saturated conditions of teeth, flux leaks through the frame and end shields causing eddy current loss in them. This loss is a component of stray load loss. Stray load loss is difficult to calculate accurately and therefore it is taken as 1 % of the output of a DC machine.

EFFICIENCY OF A DC GENERATOR:

Power flow in a DC generator is shown in figure 3.19.

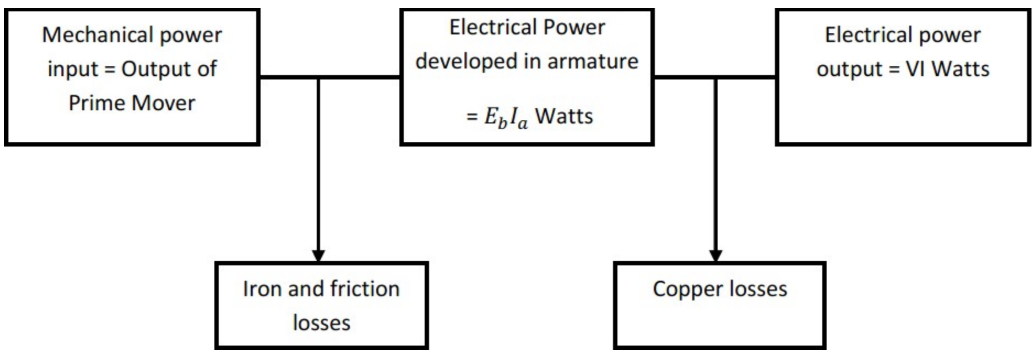


Fig. 3.19 Power flow in a DC generator

CONDITION FOR MAXIMUM EFFICIENCY

Generator output = VI;

Generator input = VI + losses.

$\text{Input} = VI + I_a^2 r_a + w_c$

If the shunt field current is negligible, then $I_a = I$

For maximum efficiency $\frac{d}{dI}(\eta) = 0$

$I_a^2 r_a = w_c$

Hence efficiency is maximum when variable loss = constant loss.

The load current corresponding to maximum efficiency is $I = \sqrt{\frac{w_c}{r_a}}$

EFFICIENCY OF DC MOTOR:

The power flow in a DC motor is shown in figure 3.20.

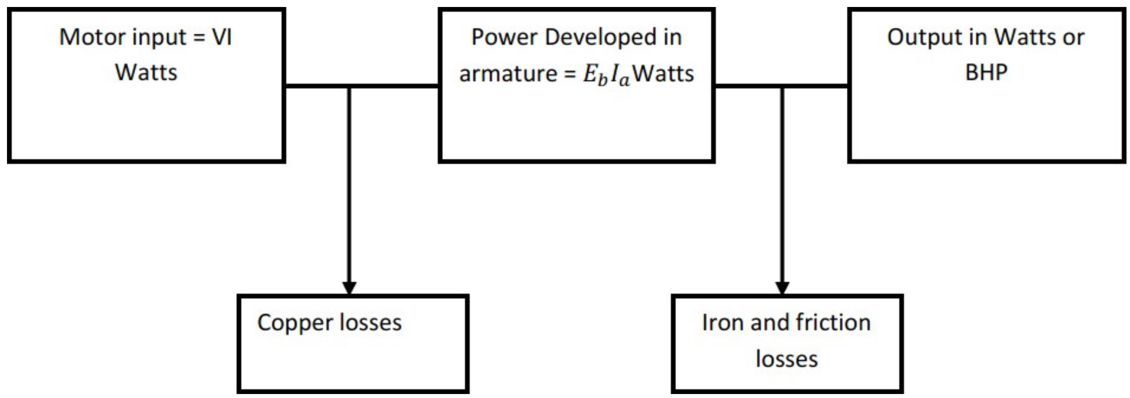


Fig. 3.20 Power flow in a DC generator

Efficiency $\eta = \dfrac{\text{Input-Losses}}{\text{Input}}$ (for Motor)

$$\eta = \frac{VI - I_a^2 r_a - w_c}{VI}$$

Efficiency is maximum when variable loss = constant loss.

Speed Control of DC Motor

DC motors are in general much more adaptable speed drives than AC motors. Speed of a DC motor can be controlled in a wide range.

$$N = \frac{E_b}{K\phi} = \frac{V - I_a r_a}{K\phi}$$

The speed equation shows that speed can be controlled by-

1. Variation of field current which varies the flux/pole (ϕ) and is known as field control.
2. Variation of armature resistance known as armature voltage control.
3. Variation of terminal voltage 'V' known as Ward Leonard method.

Field Control

For a fixed terminal voltage, $\dfrac{N_2}{N_1} = \dfrac{\phi_2}{\phi_1} = \dfrac{If_2}{If_1}$

Limitations of speed control by field control:

1. 'N' below rated speed is not possible. Because ϕ can be decreased and cannot increase.
2. $N \propto \frac{1}{\phi}$ & $T \propto \phi$ for a given armature current, this method suits for constant kW drives only where 'T' decreases if speed decreases.
3. Not suited for speed reversal.

DC Shunt Motor

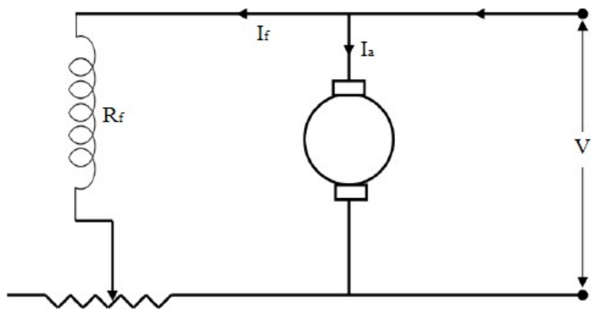


Fig. 3.21 Circuit diagram for speed control using field control method

Speed control is achieved by means of a rheostat in the field circuit as shown in the figure 3.21. The working range of the speed torque characteristics reduces with increasing speed in order for the armature current not to exceed the full load value with a weakening field.

DC Series Motor

Speed control is achieved by adjusting the field ampere turns. There are three ways for varying the field ampere turns.

A. Diverter field control

Diverter resistance R_d is connected across the field winding as shown in figure 3.22. By varying R_d the field current and hence the field ampere turns can be reduced.

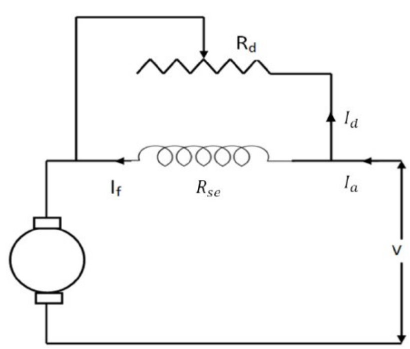


Fig. 3.22 Field diverter circuit

B. Tapped field control:

The field ampere turns are adjusted in steps by varying the number of turns included in the circuit as shown in figure 3.23. By changing number of field winding turns effective ampere turns of the field is changed thus field flux can be controlled.

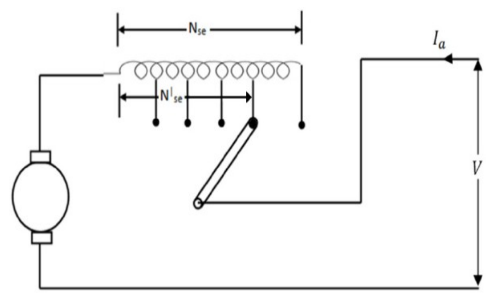


Fig. 3.23 Tapped field circuit

C. Series parallel control

In this method, the field windings are divided into two equal halves and then connected in series or parallel to control the field ampere turns as shown in figure 3.24 and 3.25respectively.

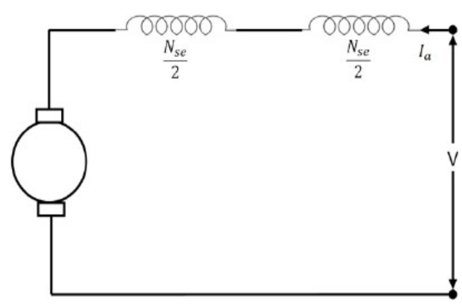


Fig. 3.24 Field circuit connected in series

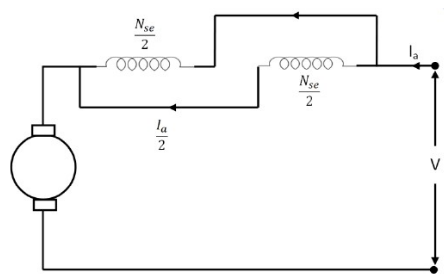


Fig. 3.25 Field circuit connected in parallel

Armature Voltage Control

In this method, applied voltage across the armature of the DC motor is varied. This method is superior to field control in the following aspects:

1. This method provides a constant torque drive.(if the ϕ and I_a are maximum, maximum torque can be obtained as $T \propto \phi I_a$)
2. Since main field ampere turns are maintained at large value, flux density distortion caused by armature reaction is limited.
3. Unlike field control scheme, speed reversal can be easily implemented.
4. This method requires a variable voltage supply which makes this method costlier.

3.12.2.1 DC Shunt Motor

Following are the armature control schemes for DC shunt motor.

1. Rheostatic control:

Here the applied armature voltage is varied by placing an adjustable resistance 'R' in series with the armature as shown in the figure 3.26. N vs T for varying 'R' is shown in figure 3.27.

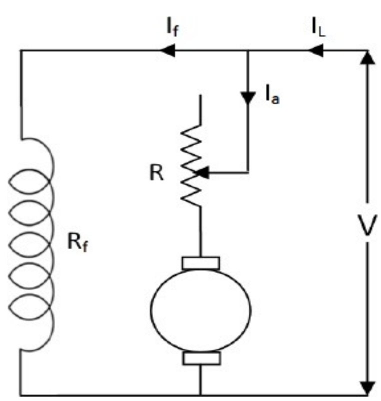


Fig. 3.27 Circuit diagram for armature resistance control

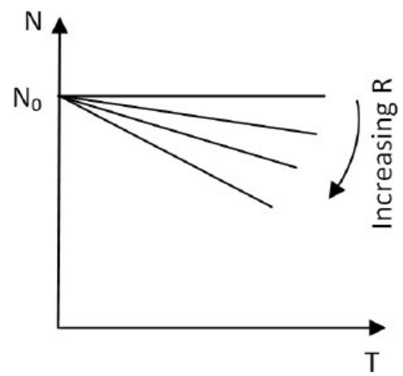


Fig. 3.28 Speed variation with torque for different resistance

Some of the limitations of the rheostatic method are:

Speeds only below rated speed

1. Range of speeds is limited because efficiency reduces drastically for large speed reductions
2. Speed regulation is poor. Because for a given resistance r_e, N varies directly with load.
3. Therefore this method is suitable for very small (fractional kW) or for short-time, intermittent slowdowns for medium sized motors.

2. Shunted armature control

In the armature rheostatic control method, the change in armature current due to change in load will affect the speed. Hence in this method the armature is shunted by an adjustable resistance as shown in figure 3.28.

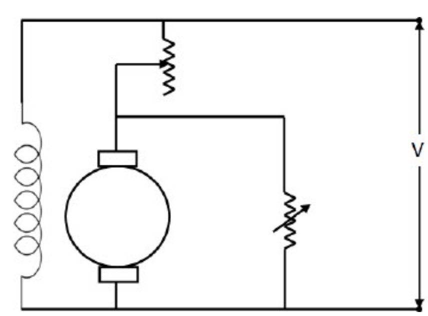


Fig. 3.28 Shunted armature control circuit

Advantages of this method are

1. Speed regulation will be better.

2. The changes in the armature current due to load will not be as effective as the armature is connected across a resistance.

Ward-Leonard Method

It is a combined armature and field control and is therefore operationally the most efficient method of speed control with a wide range. 'M_1' is the main motor whose speed control is required. The field of this motor is permanently connected across the DC supply lines. Its armature is supplied by a variable voltage derived by a Motor-Generator set. The motor M_2 act as prime mover for the generator can be AC motor or DC motor. The field of the DC generator is separately excited. The entire arrangement is shown in the figure 3.29. The reversible switch provided for the generator field makes it possible to easily reverse the generator excitation thereby reversing the voltage polarity for reversing the direction of rotation of motor. Though expensive, this arrangement can be easily adapted to feedback schemes for automatic control of speed. This method provides both constant torque and constant HP (or kW) drive.

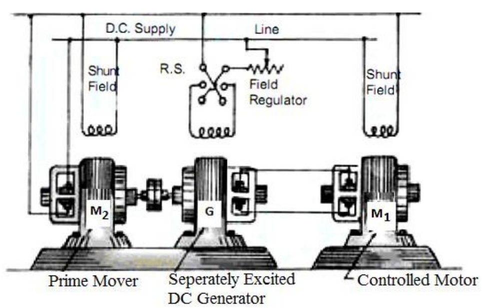


Fig. 3.29 Ward-Leonard speed control scheme

The armature and field winding of the motor are fed at maximum values at the base speed N_{base}. When armature voltage is reduced a constant torque speed control is obtained where the speed can be reduced below the base value, while the motor has full torque capability. When speed above N_{base} is required then the field is gradually weakened. The motor torque therefore reduces as its speed increases which corresponds to constant kW (or HP) drive. This is shown in the figure 3.30.

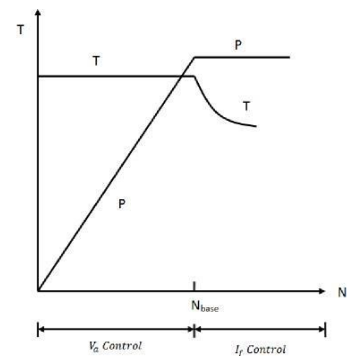


Fig. 3.30 Speed torque relation during Ward –Leonard speed control

Some of the features of the Ward Leonard system are given below:

1. As this method does not required any external resistance thus the efficiency is improved at all speeds and also when the generator emf becomes less than the back emf of the motor, the electrical power flows back from motor to generator, is converted to mechanical form and is returned to the mains via the driving AC motor.
2. Motor starts up smoothly therefore starting device is not required.
3. Reversal of speed is smoothly carried out.
4. Fine speed control from zero to rated value in both the directions are possible.

This method of speed control is used in

a. High speed elevators

b. Colliery winders

<u>Testing of DC machines</u>

Testing of DC machines can be broadly classified as

a) Direct method of Testing

b) Indirect method of testing

Direct method of testing

In this method, the DC machine is loaded directly by means of a brake applied to a water cooled pulley coupled to the shaft of the machine. The input and output are measured and efficiency is determined by $\eta = \dfrac{\text{Output}}{\text{Input}}$

It is not practically possible to arrange loads for machines of large capacity.

BRAKE TEST:

This is a direct method of testing. In this method of testing motor shaft is coupled to a Water cooled pulley which is loaded by means of weight as shown in figure 3.31.

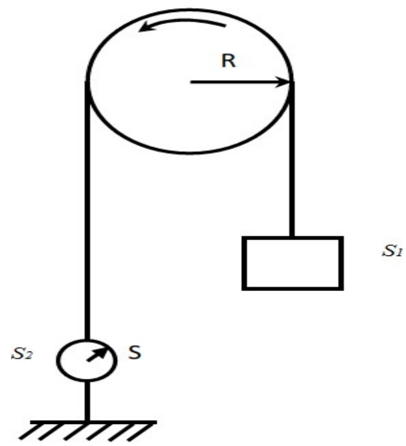


Fig. 3.31 Brake pulley arrangement for direct load test

S_1= suspended weight in kg

S_2= Reading in spring balance in kg

R = radius of pulley

n = speed in rps

V = Supply voltage

I = Full Load Current

Net pull due to friction = (S_1-S_2) kg

= 9.81 (S_1-S_2) Newton

Shaft torque = (S_1-S_2) R kg-mt

= 9.81 $(S_1 - S_2)$ R N-mt

Motor output power = T_{sh} X $2\Pi n$ Watt

= $(S_1 - S_2)$ R X $2\Pi n$ watts

Or 9.81 (S_1 — S_2)R X 2Πn watt.

Input power = VI watts

Therefore efficiency $= \eta = \frac{\text{Output}}{\text{Input}} = \frac{9.81\,(S_1 - S2)R \times 2\Pi n}{VI}$

This method of testing can be used for small motors only because for a large motor it is difficult to arrange for dissipation of heat generated at the brake.

Indirect method of testing:

In this method, the machine is not actually loaded. The losses are determined. If the losses are known, then efficiency can be determined. Swinburne's test and Hopkinson's test are commonly used on shunt motors. But, as series motor cannot be started on No-load, these tests cannot be conducted on DC series motor.

Swinburne's *Test*

For a d.c shunt motor change of speed from no load to full load is quite small. Therefore, mechanical loss can be assumed to remain same from no load to full load. Also if field current is held constant during loading, the core loss too can be assumed to remain same.

In this test, the motor is run at rated speed under *no load* condition at rated voltage. The current drawn from the supply I_{L0} and the field current I_f are recorded. Now we note that:

Input power to the motor, $P_{in} = VI_{L0}$

Cu loss in the field circuit $P_{fl} = VI_f$

Power input to the armature, $= VI_{L0} - VI_f = V(I_{L0} - I_f) = VI_{a0}$

Cu loss in the armature circuit $= I^2_{a0}\, r_a$

Gross power developed by armature $= VI_{a0} - I^2_{a0}\, r_a = \left(V - I_{a0} r_a\right) I_{a0} = E_{b0} I_{a0}$

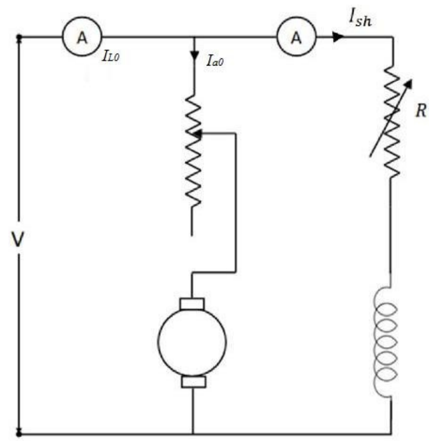


Fig. 3.32 Circuit diagram for Swinburne's Test

Since the motor is operating under no load condition, net mechanical output power is zero. Hence the gross power developed by the armature must supply the core loss and friction & windage losses of the motor. Therefore,

$$P_{core} + P_{friction} = (V - I_{a0} r_a) I_{a0} = E_{b0} I_{a0}$$

Since, both P_{core} and $P_{friction}$ for a shunt motor remains practically constant from no load to full load, the sum of these losses is called constant rotational loss i.e.,

Constant rotational loss, $P_{rot} = P_{core} + P_{friction}$

In the Swinburne's test, the constant rotational loss comprising of core and friction loss is estimated from the above equation.

After knowing the value of P_{rot} from the Swinburne's test, we can fairly estimate the efficiency of the motor at any loading condition. Let the motor be loaded such that new current drawn from the supply is I_L and the new armature current is I_a. To estimate the efficiency of the loaded motor we proceed as follows:

Input power to the motor, $P_{in} = VI_L$

Cu loss in the field circuit $P_{fl} = VI_f$

Power input to the armature, $= VI_L - VI_f = V(I_L - I_f) = VI_a$

Cu loss in the armature circuit = $I_a^2 r_a$

Gross power developed by armature = $VI_a - I_a^2 r_a = (V - I_a r_a) I_a = E_b I_a$

Net mechanical output power, $P_{net\ mech} = E_b I_a - P_{rot}$

Efficiency of the loaded motor,

$$\eta = \frac{E_b I_a - P_{rot}}{VI_L} = \frac{P_{net\ mech}}{P_{in}}$$

The estimated value of P_{rot} obtained from Swinburne's test can also be used to estimate the efficiency of the shunt machine operating as a generator.

Output power of the generator, $P_{out} = VI_L$

Cu loss in the field circuit $P_{fl} = VI_f$

Output power of the armature, $= VI_L + VI_f = VI_a$

Mechanical input power, $P_{in\ mech} = VI_a + I_a^2 r_a + P_{rot}$

Efficiency of the generator,

$$\eta = \frac{VI_L}{P_{in\ mech}} = \frac{VI_L}{VI_L + VI_f + I_a^2 r_a + P_{rot}}$$

As this test is done at no-load condition thus the power required is very less. From the test effect of armature reaction, temperature rise, commutation etc. cannot be predicted as the machine is not actually loaded.

Load Test

To assess the rating of a machine a load test has to be conducted. When the machine is loaded, certain fraction of the input is lost inside the machine and appears as heat, increasing the temperature of the machine. If the temperature rise is excessive then it affects the insulations, ultimately leading to the breakdown of the insulation and the machine. The load test gives the information about the efficiency of a given machine at any load condition. Also, it gives the temperature rise of the machine. If the temperature rise is below the permissible value for the insulation then the machine can be safely operated at that load, else the load has to be reduced. The maximum continuous load that can be delivered by the machine without exceeding the temperature rise for the insulation used, is termed as the continuous rating of the machine. Thus the load test alone can give us the proper information of the rating and also can help in the direct measurement of the efficiency.

Hopkinson's Test

Here power drawn from the supply only corresponds to no load losses of the machines, the armature physically carries any amount of current (which can be controlled with ease). Two similar DC shunt motors are mechanically coupled. Electrically these two machines are eventually connected in parallel and controlled in such a way that one machine acts as a generator and the other as motor.

Two similar (same rating) machines are connected and coupled as shown in figure 3.33. With switch is open initially, the first machine is run as a shunt motor at rated speed. It may be noted that the second machine is operating as a separately excited generator because its field winding is excited and it is driven by the first machine. The value of the voltage across the switch is either close to twice supply voltage or small voltage. In fact the voltmeter practically reads the difference of the induced voltages in the armature of the machines. In case if the voltmeter reading is high, then the armature connection of the generator should be reversed and start afresh. Now if the voltmeter is found to read small voltage then any attempt to close the switch may result into large circulating current as the armature resistances are small. By adjusting the field current I_{fg} of the generator the voltmeter reading may be adjusted to zero ($E_g \approx E_b$) and switch is now closed. Both the machines are now connected in parallel as shown in figure 3.33.

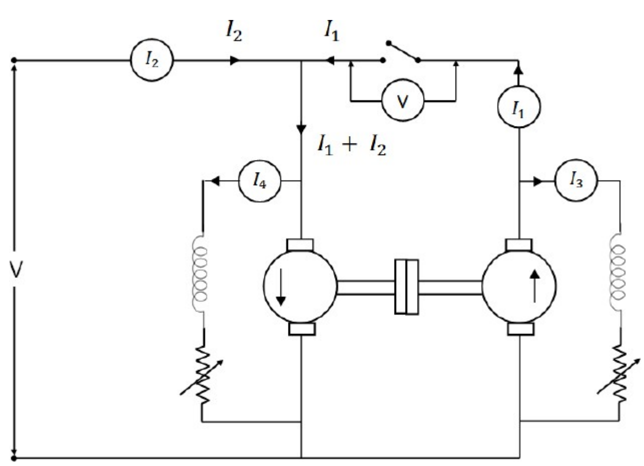


Fig. 3.33 Connection of Hopkinson's Test

After the machines are successfully connected in parallel, if the field current of generator is increased (by decreasing generator field resistance), then E_g becomes greater than E_b and both armature current of generator and motor increase, Thus by increasing field current of generator (alternatively decreasing field current of motor) one can make $E_g > E_b$ so as to make the second machine act as generator and first machine as motor. In practice, it is also required to control the field current of the motor to maintain speed constant at rated value. The interesting point to be noted here is that the armature current of generator and motor are not reflected in the supply side line. Thus current drawn from supply remains small (corresponding to losses of both the machines). The loading is sustained by the output power of the generator running the motor and vice versa. The machines can be loaded to full load current without the need of any loading arrangement.

Calculation

V= supply voltage

Motor input = $V(I_1+I_2)$

Generator output = VI_1

If it is assumed both machines have the same efficiency 'η', then,

Output of motor = η x input = η x $V(I_1+I_2)$ = input to generator

Output of generator = η x input = η x $\eta V(I_1+I_2) = \eta^2 V(I_1+I_2)$

$VI_1 = \eta^2 V(I_1+I_2)$

Therefore, $\eta = \sqrt{\dfrac{I_1}{I_1+I_2}}$

Armature copper loss in motor = $(I_1 + I_2 - I_4)^2 r_a$

Shunt field copper loss in motor = VI_4

Armature copper loss in generator = $(I_1 + I_3)^2 r_a$

Shunt field copper loss in generator = VI_3

Power drawn from supply = VI_2

Therefore stray losses = $VI_2 - [(I_1 + I_2 - I_4)^2 r_a + VI_4 + (I_1 + I_3)^2 r_a + VI_3]$ = W (say)

Stray losses/motor = $\frac{W}{2}$

Therefore for generator

Total losses= $\left(I_1+I_2\right)^2 r_a+VI_3+\frac{W}{2}=W_g$

Output = VI_1, therefore

$\eta_{generator}=\frac{VI_1}{VI_1+W_g}=\frac{output}{output+losses}$

For motor,

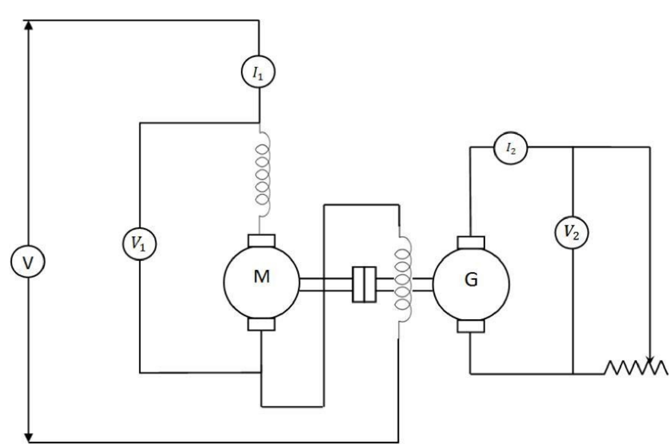
Total losses = $\left(I_1+I_2-I_4\right)^2 r_a+VI_4+\frac{W}{2}=W_m$

Input to motor = $V(I_1+I_2)$

Therefore $\eta_{motor}=\frac{V(I_1+I_2)-W_m}{V(I_1+I_2)}$

Field's Test

Figure 3.34 shows the circuit for fields test. This test is applicable to two similar series motor. One of the machine runs as a motor and drives a generator whose output is wasted in a variable load 'R'. Both machine field coils are in series and both run at same speed so that iron and friction losses are made equal.



3.34 Circuit diagram for Field's test on DC series motor

Load resistance 'R' is varied till the motor current reaches its full load value.

V = Supply voltage

I_1 = Motor current

V_2 = Generator terminal voltage

I_2 = Load current

Input = VI_1 and output = V_2I_2

R_a and R_{se} = hot resistances.

Total losses in the set $W_t = VI_1 - V_2I_2$

Armature and Field copper losses $W_c = (R_a + 2\, r_{se})\, I_1^2 + I_a^2R_a$

Stray losses for the set = $W_t - W_c$

Stray losses per machine $W_s = \dfrac{W_t - W_c}{2}$

Motor efficiency :

Input=V_1I_1

Losses=$\left(R_a + R_{se}\right)I_1^2 + W_s = W_m$ (say)

$$\eta_{motor} = \frac{VI_1 + W_m}{VI_1}$$

Generator efficiency: η of generator is of little use, because its field winding is separately excited

Generator output = VI_2

Field copper loss = $I_1^2 r_{se}$

Armature copper loss = $I_1^2 r_a$

Total losses = $I_1^2 r_{se} + I_1^2 r_a + W_s = W_g$ (say)

$$\eta_{generator} = \frac{VI_2}{V_2 + W_g}$$

3.12.2.5Reatardation Test

This method is applicable to shunt motors and generators and is used for finding the stray losses. If armature and shunt copper losses are known for a given load, efficiency can be calculated. The circuit is shown in figure 3.35.

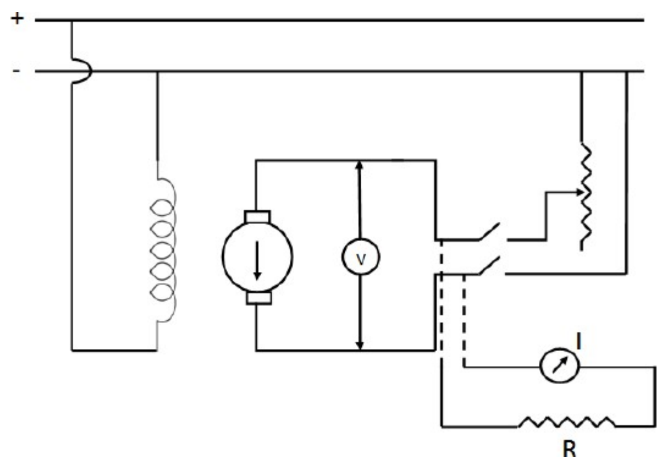


Fig. 3.35 Circuit diagram for Retardation test on DC motor

Machine is speeded up slightly beyond its rated speed and then supply is cut off from the armature while keeping the field excited. Armature will slow down and its kinetic energy is needed to meet rotational losses. i.e., friction and windage losses.

Kinetic energy of the armature = $\frac{1}{2} I\omega^2$

I = Moment of inertia of the armature

ω = Angular velocity.

Rotational losses;

N = Rate of loss of K.E.

Rate of loss of Kinetic energy W = $\frac{d}{dt}\left[\frac{1}{2} I\omega^2\right] = I\omega \frac{d\omega}{dt}$

Two quantities need to be known

(i) Moment of Inertia 'I

(ii) $\frac{d\omega}{dt}$ or $\frac{dN}{dt}$ (because $\omega \propto N$)

(i) Finding $\frac{d\omega}{dt}$:

The voltmeter "V" in the circuit shown in Fig. 3.35 is used as per speed indicator by suitably grading it because $E \propto N$. Then the supply is cut off, the armature speed and hence voltmeter reading falls. Voltage and time at different interval are noted and a curve is drawn between time and speed as shown in fig. 3.36.

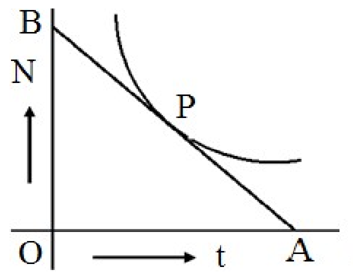


Fig. 3.36 Change of speed with time

In the fig. 3.36 AB- tangent drawn at P

Therefore $\frac{dN}{dt} = \frac{OB(rpm)}{OA(sec)}$

$$W = I \times \omega \times \frac{d\omega}{dt}$$

$$\omega = \frac{2\pi N}{60}$$

$$W = I\left(\frac{2\pi N}{60}\right)\frac{d}{dt}\left(\frac{2\pi N}{60}\right)$$

$$W = \left(\frac{2\pi}{60}\right)^2 .I.N.\frac{dN}{dt}$$

(ii) Finding Moment of Inertia "I":

There are two methods of finding the moment of inertia 'I'

(a) I is calculated:

(i) Slowing down curve with armature alone is calculated.

(ii) A fly wheel is keyed to the shaft and the curve is drawn again

For any given speed, $\frac{dN}{dt}$ and $\frac{dN}{dt_1}$ are determined as before.

Therefore $W=\left(\frac{2\pi}{60}\right)^2 .I.N.\frac{dN}{dt_1}$ 1st case

$W=\left(\frac{2\pi}{60}\right)^2 \left(I+I_1\right)N.\frac{dN}{dt_2}$ 2nd Case

The two cases are equal because losses in two cases will be almost same.

$I\frac{dN}{dt}=\left(I+I_1\right)\frac{dN}{dt}.\frac{I+I1}{I}\left(\frac{dN}{dt_2}\right)=\frac{dN}{dt_1}$

$\frac{I+I1}{I}=\frac{dN}{dt_2}$

$I=I_1\times\frac{t_2}{t_1-t_2}$

(b) I is eliminated:

In this method, time taken to slow down is noted with armature alone and then a retarding torque is applied electrically i.e., a non-inductive resistance is connected to the armature.

The additional loss is $I_a^2\left(R_a+R\right)orVI_a$

Let W^1 be the power then

$W=\left(\frac{2\pi}{60}\right)^2 IN.\frac{dN}{dt_1}$

$W+W^1=\left(\frac{2\pi}{60}\right)^2 IN.\frac{dN}{dt_2}$

$\frac{dN}{dt_1}$ = rate of change of speed without electrical load

$\frac{dN}{dt_2}$ = rate of change of speed with electrical load

$\frac{W+W^1}{W}=-\frac{\frac{dN}{dt_2}}{\frac{dN}{dt_1}}$

or, $W=W^1\times\frac{dt_2}{dt_1-dt_2}$

or $W=W^1\times\frac{t_2}{t_1-t_2}$

Module IV

[THREE PHASE TRANSFORMER]

TOPICS

Three Phase Transformers: Constructional features of three phase transformers – three phase connection of transformers (Dd0, Dd6, Yy0, Yy6, Dy1, Dy11, Yd1, Yd11, zigzag), Scott connection, open delta connection, three phase to six phase connection, oscillating neutral, tertiary winding, three winding transformer, equal and unequal turns ratio, parallel operation, load sharing. Distribution transformers, all day efficiency, Autotransformers, saving of copper, applications, tap- changing transformers, cooling of transformers.

[Topics are arranged as per above sequence]

Three Phase Transformers

Introduction

Electric power is generated in generating stations, using three phase alternators at 11 KV. This voltage is further stepped up to 66 KV, 110 KV, 230 KV or 400 KV using 3 phase power transformers and power is transmitted at this high voltage through transmission lines. At the receiving substations, these high voltages are stepped down by 3 phase transformers to 11 KV. This is further stepped down to 400 volts at load centers by means of distribution transformers. For generation, transmission and distribution, 3 phase system is economical. Therefore 3 phase transformers are very essential for the above purpose. The sectional view of a 3 phase power transformer is shown in Fig.4.1.

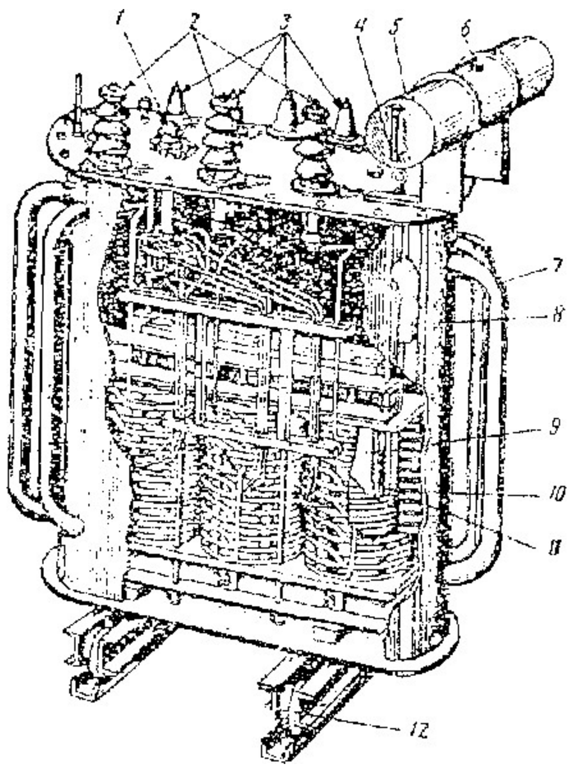


Fig. 4.1 100 KVA oil immersed power transformer

1. Tap-changer switch handle
2. Porcelain-bushing insulator (For high voltage)
3. Bushing insulators (For low voltages)
4. Oil gauge

5. Oil tank
6. Breather plug
7. Cooling pipes
8. Tank front wall
9. Core,
10. High voltage winding
11. Low voltage winding
12. Wheels or rollers.

Construction of Three phase Transformer

Three phase transformers comprise of three primary and three secondary windings. They are wound over the laminated core as we have seen in single phase transformers. Three phase transformers are also of core type or shell type as in single phase transformers. The basic principle of a three phase transformer is illustrated in fig 4.2 in which the primary windings and secondary windings of three phases are shown. The primary windings can be inter connected in star or delta and put across three phase supply.

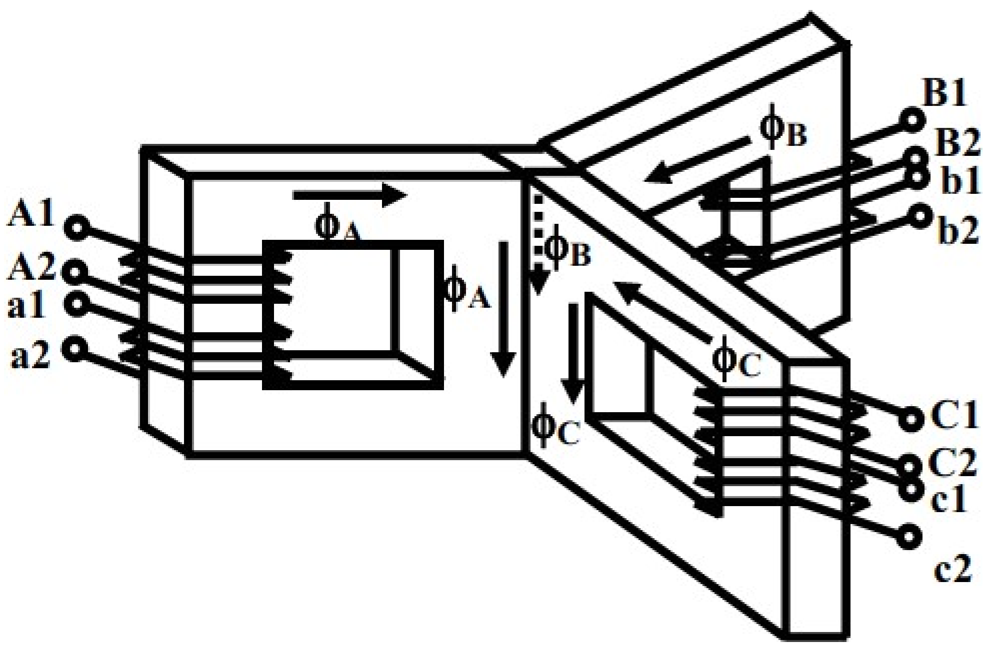


Fig. 4.2 3-phase core-type Transformer

The three cores are 120° apart and their unwound limbs are shown in contact with each other. The center core formed by these three limbs, carries the flux produced by the three phase currents I_R, I_Y and I_B. As at any instant $I_R+I_y+I_B=0$, the sum of three fluxes (flux in the center limb) is also zero.

Therefore it will make no difference if the common limb is removed. All the three limbs are placed in one plane in case of a practical transformer as shown in fig 4.3.

The core type transformers are usually wound with circular cylindrical coils. The construction and assembly of laminations and yoke of a three phase core type transformer is shown in fig 4.4 one method of arrangement of windings in a three phase transformer is shown.

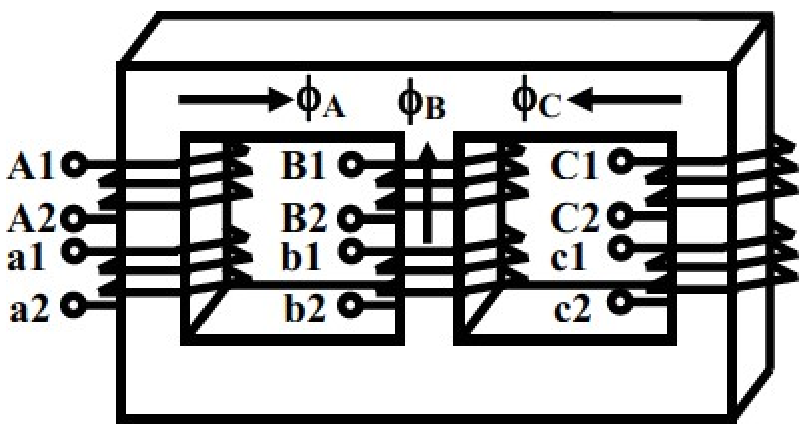


Fig. 4.3 A practical core type three phase transformer

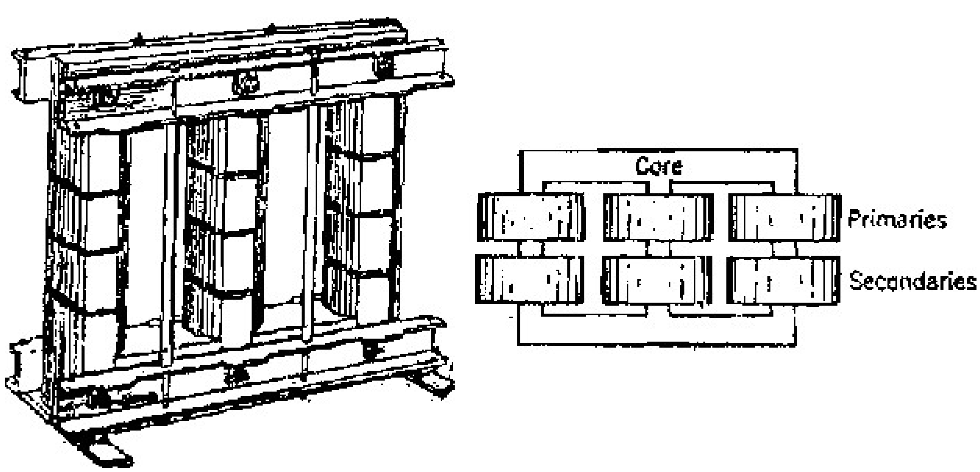


Fig. 4.4 Core type transformer windings and construction

In the other method the primary and secondary windings are wound one over the other in each limb. The low-tension windings are wound directly over the core but are, of course, insulated for it. The high tension windings are wound over the low— tension windings and adequate insulation is provided between the two windings.

The primary and secondary windings of the three phase transformer can also be interconnected as star or delta.

Three Phase Transformer connections:-

The identical single phase transformers can be suitably inter-connected and used instead of a single unit 3—phase transformer. The single unit 3 phase transformer is housed in a single tank. But the transformer bank is made up of three separate single phase transformers each with its own, tanks and bushings. This method is preferred in mines and high altitude power stations because transportation becomes easier. Bank method is adopted also when the voltage involved is high because it is easier to provide proper insulation in each single phase transformer.

As compared to a bank of single phase transformers, the main advantages of a single unit 3-phase transformer are that it occupies less floor space for equal rating, less weight costs about 20% less and further that only one unit is to be handled and connected.

There are various methods available for transforming 3 phase voltages to higher or lower 3 phase voltages. The most common connections are (i) star — star (ii) Delta—Delta (iii) Star —Delta (iv) Delta — Star.

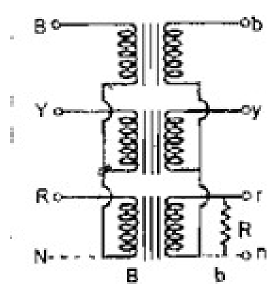


Fig 4.5 Star-star connection

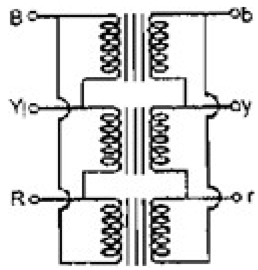


Fig. 4.6 Delta-delta connection

The star-star connection is most economical for small, high voltage transformers because the number of turns per phase and the amount of insulation required is minimum (as phase voltage is only 1/3 of line voltage. In fig. 4.5 a bank of three transformers connected in star on both the primary and the secondary sides is shown. The ratio of line voltages on the primary to the secondary sides is the same as a transformation ratio of single phase transformer.

The delta— delta connection is economical for large capacity, low voltage transformers in which insulation problem is not a serious one. The transformer connection are as shown in fig. 4.6.

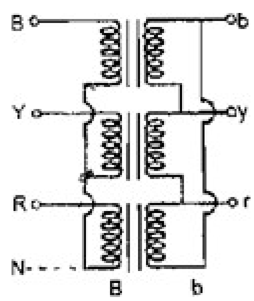


Fig. 4.7 Star-delta connection

The main use of star-delta connection is at the substation end of the transmission line where the voltage is to be stepped down. The primary winding is star connected with grounded neutral as shown in Fig. 4.7. The ratio between the secondary and primary line voltage is 1/3 times the transformation ratio of each single phase transformer. There is a 30° shift between the primary and secondary line voltages which means that a star-delta transformer bank cannot be paralleled with either a star-star or a delta-delta bank.

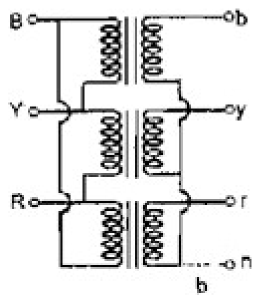


Fig. 4.8 Delta-star connection

Delta-Star connection is generally employed where it is necessary to step up the voltage. The connection is shown in fig. 4.8. The neutral of the secondary is grounded for providing 3-phase, 4-wire service. The connection is very popular because it can be used to serve both the 3-phase power equipment and single phase lighting circuits.

Vector Group of 3-phase transformer

The secondary voltages of a 3-phase transformer may undergo a *phase shift* of either +30° leading or -30° lagging or 0° i.e, no phase shift or 180° reversal with respective line or phase to neutral voltages. On the name plate of a three phase transformer, the vector group is mentioned. Typical representation of the vector group could be Ydl or Dy 11 etc. The first capital latter Y indicates that the primary is connected in star and the second lower case latter d indicates delta connection of the secondary side. The third numerical figure conveys the angle of phase shift based on *clock convention.* The minute hand is used to represent the primary phase to neutral voltage and always shown to occupy the position 12. The hour hand represents the secondary phase to neutral voltage and may, depending upon phase shift, occupy position other than 12 as shown in the figure 4.9. The angle between two consecutive numbers on the clock is 30°.

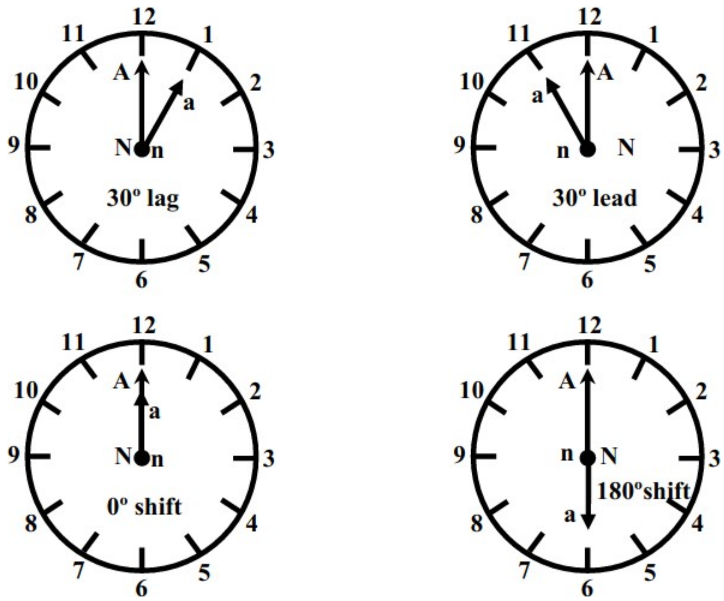


Fig. 4.9 Clock convention representing vector groups

Delta/delta (Dd0, Dd6) connection

The connection of Dd0 is shown in fig. 4.10 and the voltages on primary and secondary sides is also shown on the phasor diagram. The phase angle difference between the phase voltage of high voltage side and low voltage side is zero degree (0°).

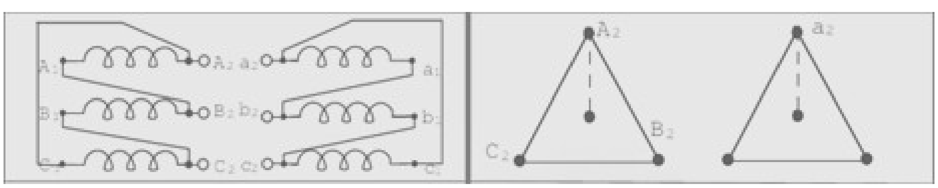


Fig 4.10 Dd0 connection and phasor diagram

The connection of Dd6 is shown in fig. 4.11 and the voltages on primary and secondary sides is also shown on the phasor diagram. The phase angle difference between the phase voltage of high voltage side and low voltage side is 180°.

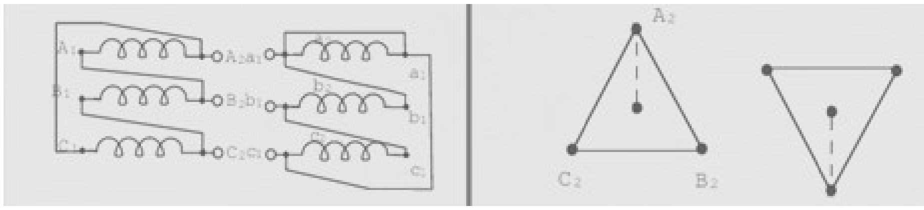


Fig 4.11 Dd6 connection and phasor diagram

This connection proves to be economical for large low voltage transformers as it increases number of turns per phase. Primary side line voltage is equal to secondary side line voltage. Primary side phase voltage is equal to secondary side phase voltage. There is no phase shift between primary and secondary voltages for Dd0 connection. There is 180° phase shift between primary and secondary voltages for Dd6 connection.

Advantages

- **Sinusoidal Voltage at Secondary:** In order to get secondary voltage as sinusoidal, the magnetizing current of transformer must contain a third harmonic component. The delta connection provides a closed path for circulation of third harmonic component of current. The flux remains sinusoidal which results in sinusoidal voltages.

- **Suitable for Unbalanced Load:** Even if the load is unbalanced the three phase voltages remains constant. Thus it suitable for unbalanced loading also.

- **Carry 58% Load if One Transfer is Faulty in Transformer Bank**: If there is bank of single phase transformers connected in delta-delta fashion and if one of the transformers is disabled then the supply can be continued with remaining tow transformers of course with reduced efficiency.

- **No Distortion in Secondary Voltage**: there is no any phase displacement between primary and secondary voltages. There is no distortion of flux as the third harmonic component of magnetizing current can flow in the delta connected primary windings without flowing in the line wires .there is no distortion in the secondary voltages.

- **Economical for Low Voltage:** Due to delta connection, phase voltage is same as line voltage hence winding have more number of turns. But phase current is (1/√3) times the line current. Hence the cross-section of the windings is very less. This makes the connection economical for low voltages transformers.
- **Reduce Cross section of Conductor:** The conductor is required of smaller Cross section as the phase current is 1/√3 times of the line current. It increases number of turns per phase and reduces the necessary cross sectional area of conductors thus insulation problem is not present.
- **Absent of Third Harmonic Voltage:** Due to closed delta, third harmonic voltages are absent.
- The absence of star or neutral point proves to be advantageous in some cases.

Disadvantages

- Due to the absence of neutral point it is not suitable for three phase four wire system.
- More insulation is required and the voltage appearing between windings and core will be equal to full line voltage in case of earth fault on one phase.

Application

- Suitable for large, low voltage transformers.
- This Type of Connection is normally uncommon but used in some industrial facilities to reduce impact of SLG faults on the primary system
- It is generally used in systems where it need to be carry large currents on low voltages and especially when continuity of service is to be maintained even though one of the phases develops fault.

Star/star (YyO, Yy6) connection

This is the most economical one for small high voltage transformers. Insulation cost is highly reduced. Neutral wire can permit mixed loading. Triplen harmonics are absent in the lines. These triplen harmonic currents cannot flow, unless there is a neutral wire. This connection produces oscillating neutral. Three phase shell type units have large triplen harmonic phase voltage. However three phase core type transformers work satisfactorily. A tertiary mesh connected winding may be required to stabilize the oscillating neutral due to third harmonics in three phase banks.

The connection of Yy0 is shown in fig. 4.12 and the voltages on primary and secondary sides is also shown on the phasor diagram. The phase angle difference between the phase voltage of high voltage side and low voltage side is zero degree (0°).

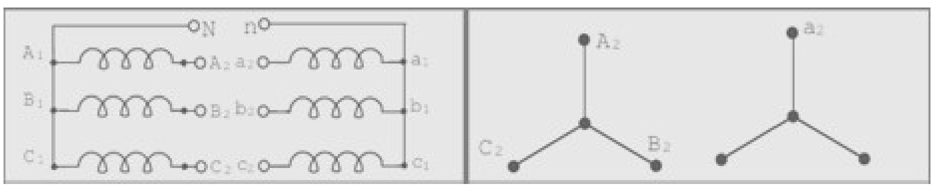


Fig .4.12 Yy0 connection and phasor diagram

The connection of Yy6 is shown in fig. 4.13 and the voltages on primary and secondary sides is also shown on the phasor diagram. The phase angle difference between the phase voltage of high voltage side and low voltage side is 180°.

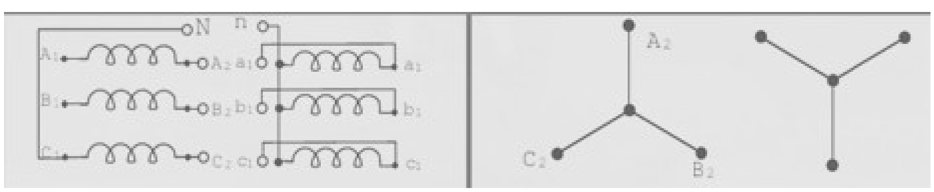


Fig 4.13. Yy6 connection and phasor diagram

- In Primary Winding Each Phase is120°electrical degrees out of phase with the other two phases.
- In Secondary Winding Each Phase is120°electrical degrees out of phase with the other two phases.

- Each primary winding is magnetically linked to one secondary winding through a common core leg. Sets of windings that are magnetically linked are drawn parallel to each other in the vector diagram. In the Y-Y connection, each primary and secondary winding is connected to a neutral point.
- The neutral point may or may not be brought out to an external physical connection and the neutral may or may not be grounded.

Advantages of Y-y connection

- **No Phase Displacement:** The primary and secondary circuits are in phase; i.e., there are no phase angle displacements introduced by the Y-Y connection. This is an important advantage when transformers are used to interconnect systems of different voltages in a cascading manner. For example, suppose there are four systems operating at 800, 440, 220, and 66 kV that need to be interconnected. Substations can be constructed using Y-Y transformer connections to interconnect any two of these voltages. The 800 kV systems can be tied with the 66 kV systems through a single 800 to 66 kV transformation or through a series of cascading transformations at 440,220 and 66 kV.
- **Required Few Turns for winding:** Due to star connection, phase voltages is (1/√3) times the line voltage. Hence less number of turns is required. Also the stress on insulation is less. This makes the connection economical for small high voltage purposes.
- **Required Less Insulation Level**: If the neutral end of a Y-connected winding is grounded, then there is an opportunity to use reduced levels of insulation at the neutral end of the winding. A winding that is connected across the phases requires full insulation throughout the winding.
- **Handle Heavy Load:** Due to star connection, phase current is same as line current. Hence windings have to carry high currents. This makes cross section of the windings high. Thus the windings are mechanically strong and windings can bear heavy loads and short circuit current.
- **Use for Three phases Four Wires System:**As neutral is available, suitable for three phases four wire

system.

- **Eliminate Distortion in Secondary Phase Voltage:** The connection of primary neutral to the neutral of generator eliminates distortion in the secondary phase voltages by giving path to triple frequency currents toward to generator.
- **Sinusoidal voltage on secondary side:** Neutral give path to flow Triple frequency current to flow Generator side thus sinusoidal voltage on primary will give sinusoidal voltage on secondary side.
- **Used as Auto Transformer:** A Y-Y transformer may be constructed as an autotransformer, with the possibility of great cost savings compared to the two-winding transformer construction.
- **Better Protective Relaying:** The protective relay settings will be protecting better on the line to ground faults when the Y-Y transformer connections with solidly grounded neutrals are applied.

Disadvantages

- **The Third harmonic issue:** The voltages in any phase of a Y-Y transformer are 1200 apart from the voltages in any other phase. However, the third-harmonic components of each phase will be in phase with each other. Nonlinearities in the transformer core always lead to generation of third harmonic. These components will add up resulting in large (can be even larger than the fundamental component) third harmonic component.
- **Overvoltage at Lighting Load:** The presence of third (and other zero-sequence) harmonics at an ungrounded neutral can cause overvoltage conditions at light load. When constructing a Y-Y transformer using single-phase transformers connected in a bank, the measured line-to-neutral voltages are not 57.7% of the system phase-to-phase voltage at no load but are about 68% and diminish very rapidly as the bank is loaded. The effective values of voltages at different frequencies combine by taking the square root of the sum of the voltages squared. With sinusoidal phase-to-phase voltage, the third-harmonic component of the phase-to-neutral

voltage is about 60%.

- **Voltage drop at Unbalance Load:** There can be a large voltage drop for unbalanced phase-to-neutral loads. This is caused by the fact that phase-to-phase loads cause a voltage drop through the leakage reactance of the transformer whereas phase-to-neutral loads cause a voltage drop through the magnetizing reactance, which is 100 to 1000 times larger than the leakage reactance.
- **Overheated Transformer Tank:** Under certain circumstances, a Y-Y connected three-phase trans- can produce severe tank overheating that can quickly destroy the transformer. This usually occurs with an open phase on the primary circuit and load on the secondary.
- **Over Excitation of Core in Fault Condition:** If a phase-to-ground fault occurs on the primary circuit with the primary neutral grounded, then the phase-to-neutral voltage on the un faulted phases increases to 173% of the normal voltage. This would almost certainly result in over excitation of the core, with greatly increased magnetizing currents and core losses
- If the neutrals of the primary and secondary are both brought out, then a phase-to-ground fault on the secondary circuit causes neutral fault current to flow in the primary circuit. Ground protection re- laying in the neutral of the primary circuit may then operate for faults on the secondary circuit
- **Neutral Shifting:** If the load on the secondary side unbalanced then the performance of this connection is not satisfactory then the shifting of neutral point is possible. To prevent this, star point of the primary is required to be connected to the star point of the generator.
- **Distortion of Secondary voltage:** Even though the star or neutral point of the primary is earthed, the third harmonic present in the alternator voltage may appear on the secondary side. This causes distortion in the secondary phase voltages.
- **Over Voltage at Light Load:** The presence of third (and other zero-sequence) harmonics at an ungrounded neutral can cause overvoltage conditions at light load.

- **Difficulty in coordination of Ground Protection:** In Y-Y Transformer, a low-side ground fault causes primary ground fault current, making coordination more difficult.
- **Increase Healthy Phase Voltage under Phase to ground Fault:** If a phase-to-ground fault occurs on the primary circuit with the primary neutral grounded, then the phase-to-neutral voltage on the UN faulted phase's increases to 173% of the normal voltage. If the neutrals of the primary and secondary are both brought out, then a phase-to-ground fault on the secondary circuit causes neutral fault current to flow in the primary circuit.
- **Trip the T/C in Line-Ground Fault:** All harmonics will propagate through the transformer, zero-sequence current path is continuous through the transformer, one line-to-ground fault will trip the transformer.
- **Suitable for Core Type Transformer:** The third harmonic voltage and current is absent in such type of connection with three phase wire system or shell type of three phase units, the third harmonic phase voltage may be high. This type of connection is more suitable for core type transformers.

Application

- This Type of Transformer is rarely used due to problems with unbalanced loads.
- It is economical for small high voltage transformers as the number of turns per phase and the amount of insulation required is less.

Star/Delta connection(Yd1/Yd11)

There is a +30 Degree or -30 Degree Phase Shift between Secondary Phase Voltage to Primary Phase Voltage. The connection of Yd1 is shown in fig. 4.14 and the voltages on primary and secondary sides is also shown on the phasor diagram. The phase angle difference between the phase voltage of high voltage side and low voltage side is -30°.

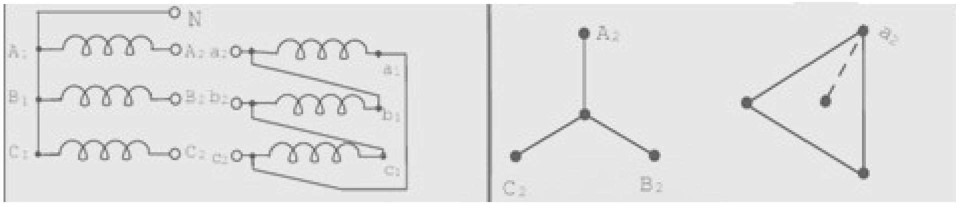


Fig 4.14. Yd1 connection and phasor diagram

The connection of Yd11 is shown in fig. 4.15 and the voltages on primary and secondary sides is also shown on the phasor diagram. The phase angle difference between the phase voltage of high voltage side and low voltage side is 30°.

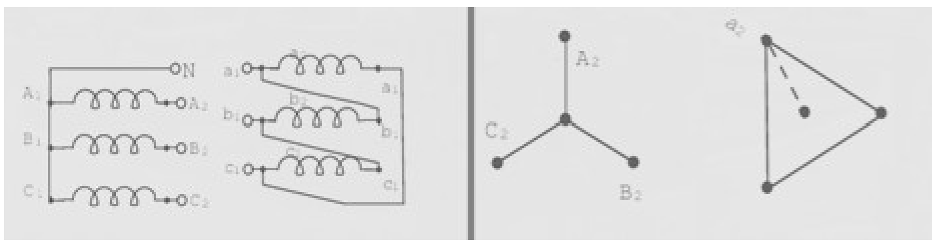


Fig 4.15. Yd11 connection and phasor diagram

Advantages

- The primary side is star connected. Hence fewer numbers of turns are required. This makes the connection economical for large high voltage step down power transformers.
- The neutral available on the primary can be earthed to avoid distortion.
- The neutral point allows both types of loads (single phase or three phases) to be met.
- Large unbalanced loads can be handled satisfactory.
- The Y-D connection has no problem with third harmonic components due to circulating currents inD. It is also more stable to unbalanced loads since the D partially redistributes any imbalance that occurs.
- The delta connected winding carries third harmonic current due to which potential of neutral point is stabilized. Some saving in cost of insulation is achieved if HV side is star connected. But in practice the HV side is normally connected in delta so that the three phase loads like motors and single phase loads like lighting loads can be supplied by LV side using three phase

four wire system.

- **As Grounding Transformer:** In Power System Mostly grounded Y- Δ transformer is used for no other purpose than to provide a good ground source in ungrounded Delta system.

Disadvantages

- In this type of connection, the secondary voltage is not in phase with the primary. Hence it is not possible to operate this connection in parallel with star-star or delta-delta connected transformer.
- One problem associated with this connection is that the secondary voltage is shifted by 30^0 with respect to the primary voltage. This can cause problems when paralleling 3-phase transformers since transformers secondary voltages must be in-phase to be paralleled. Therefore, we must pay attention to these shifts.
- If secondary of this transformer should be paralleled with secondary of another transformer without phase shift, there would be a problem

Application

- It is commonly employed for power supply transformers.
- This type of connection is commonly employed at the substation end of the transmission line. The main use with this connection is to step down the voltage. The neutral available on the primary side is grounded. It can be seen that there is phase difference of 30° between primary and secondary line voltages.
- Commonly used in a step-down transformer, Y connection on the HV side reduces insulation costs the neutral point on the HV side can be grounded, stable with respect to unbalanced loads. As for example, at the end of a transmission line. The neutral of the primary winding is earthed. In this system, line voltage ratio is $1/\sqrt{3}$ Times of transformer turn-ratio and secondary voltage lags behind primary voltage by 30°. Also third harmonic currents flows in

to give a sinusoidal flux.

Delta-star connection (Dy1/Dy11)

In this type of connection, the primary connected in delta fashion while the secondary current is connected in star. There is s +30 Degree or -30 Degree Phase Shift between Secondary Phase Voltage to Primary Phase Voltage.

The connection of Dy1 is shown in fig. 4.16 and the voltages on primary and secondary sides is also shown on the phasor diagram. The phase angle difference between the phase voltage of high voltage side and low voltage side is -30°.

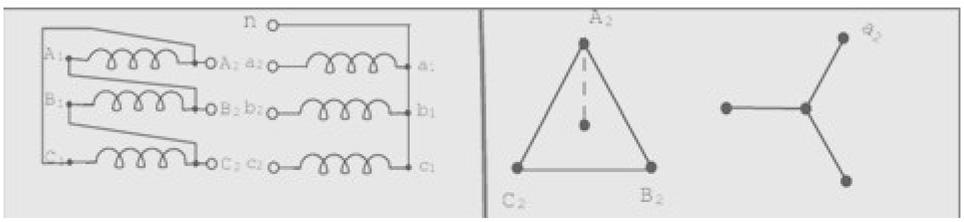


Fig 4.16. Dy1 connection and phasor diagram

The connection of Dy11 is shown in fig. 4.17 and the voltages on primary and secondary sides is also shown on the phasor diagram. The phase angle difference between the phase voltage of high voltage side and low voltage side is 30°.

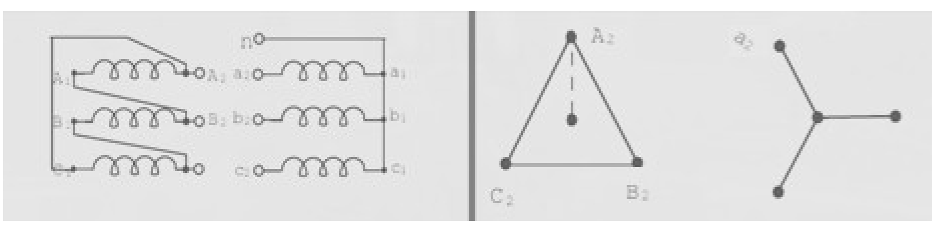


Fig 4.17. Dy11 connection and phasor diagram

Advantages

- **Cross section area of winding is less at Primary side**: On primary side due to delta connection winding cross-section required is less.

- **Used at Three phase four wire System:** On secondary side, neutral is available, due to which it can be used for 3-phase, 4 wire supply system.
- **No distortion of Secondary Voltage:** No distortion due to third harmonic components.
- **Handled large unbalanced Load:** Large unbalanced loads can be handled without any difficulty.
- **Grounding Isolation between Primary and Secondary:** Assuming that the neutral of the Y-connected secondary circuit is grounded, a load connected phase-to-neutral or a phase-to-ground fault produces two equal and opposite currents in two phases in the primary circuit without any neutral ground current in the primary circuit. Therefore, in contrast with the Y-Y connection, phase-to-ground faults or current unbalance in the secondary circuit will not affect ground protective relaying applied to the primary circuit. This feature enables proper coordination of protective devices and is a very important design consideration.
- The neutral of the Y grounded is sometimes referred to as a grounding bank, because it provides a local source of ground current at the secondary that is isolated from the primary circuit.
- **Harmonic Suppression:** The magnetizing current must contain odd harmonics for the induced voltages to be sinusoidal and the third harmonic is the dominant harmonic component. In a three-phase system the third harmonic currents of all three phases are in phase with each other because they are zero-sequence currents. In the Y-Y connection, the only path for third harmonic current is through the neutral. In the Δ -Y connection, however, the third harmonic currents, being equal in amplitude and in phase with each other, are able to circulate around the path formed by the Δ connected winding. The same thing is true for the other zero-sequence harmonics.
- **Grounding Bank:** It provides a local source of ground current at the secondary that is isolated from the primary circuit. For suppose an ungrounded generator supplies a simple radial system

through Δ-Y transformer with grounded Neutral at secondary as shown Figure. The generator can supply a single-phase-to-neutral load through the -grounded Y transformer.

Disadvantages

- In this type of connection, the secondary voltage is not in phase with the primary. Hence it is not possible to operate this connection in parallel with star-star or delta-delta connected transformer.

- One problem associated with this connection is that the secondary voltage is shifted by 30^0 with respect to the primary voltage. This can cause problems when paralleling 3-phase transformers since transformers secondary voltages must be in-phase to be paralleled. Therefore, we must pay attention to these shifts.

- If secondary of this transformer should be paralleled with secondary of another transformer without phase shift, there would be a problem.

Application

- **Commonly used in a step-up transformer:** As for example, at the beginning of a HT transmission line. In this case neutral point is stable and will not float in case of unbalanced loading. There is no distortion of flux because existence of a Δ -connection allows a path for the third-harmonic components. The line voltage ratio is √3 times of transformer turn-ratio and the secondary voltage leads the primary one by 30°. In recent years, this arrangement has become very popular for distribution system as it provides 3- Ø, 4-wire system.

- **Commonly used in commercial, industrial, and high-density residential locations**: To supply three-phase distribution systems. An example would be a distribution transformer with a delta primary, running on three 11kV phases with no neutral or earth required, and a star (or wye) secondary providing a 3-phase supply at 400 V, with the domestic voltage of 230 available between each phase and an earthed neutral point.

- **Used as Generator Transformer:** The Δ-Y transformer connection is used universally for connecting generators to transmission systems.

Delta-zigzag and Star zigzag connections (Dz0/Dz6 & Yz1/Yz6) –

The connection of Dz0 is shown in fig. 4.18 and the voltages on primary and secondary sides is also shown on the phasor diagram. The phase angle difference between the phase voltage of high voltage side and low voltage side is 0°.

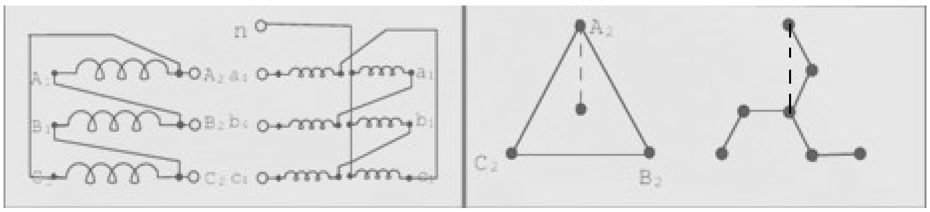


Fig 4.18. Dz0 connection and phasor diagram

The connection of Dz6 is shown in fig. 4.19 and the voltages on primary and secondary sides is also shown on the phasor diagram. The phase angle difference between the phase voltage of high voltage side and low voltage side is 180°.

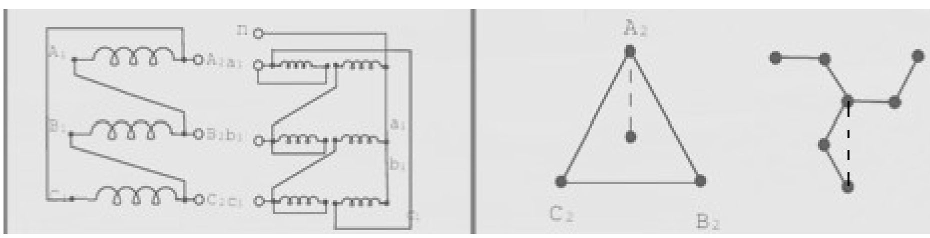


Fig 4.19. Dz6 connection and phasor diagram

The connection of Yz1 is shown in fig. 4.20 and the voltages on primary and secondary sides is also shown on the phasor diagram. The phase angle difference between the phase voltage of high voltage side and low voltage side is -30°.

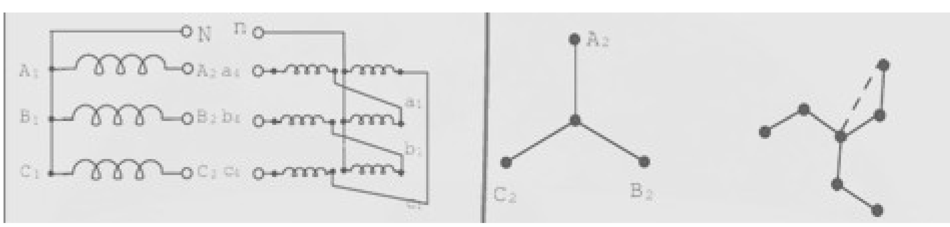


Fig 4.20. Yz1 connection and phasor diagram

The connection of Yz11 is shown in fig. 4.21 and the voltages on primary and secondary sides is also shown on the phasor diagram. The phase angle difference between the phase voltage of high voltage side and low voltage side is 30°.

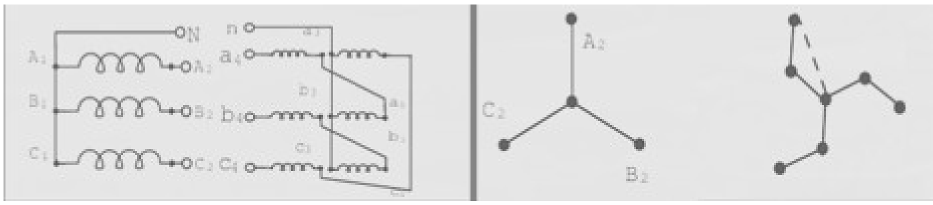


Fig 4.22 Yz11 connection and phasor diagram

- These connections are employed where delta connections are weak. Interconnection of phases in zigzag winding effects a reduction of third harmonic voltages and at the same time permits unbalanced loading.
- This connection may be used with either delta connected or star connected winding either for step-up or step-down transformers. In either case, the zigzag winding produces the same angular displacement as a delta winding, and at the same time provides a neutral for earthing purposes.
- The amount of copper required from a zigzag winding in 15% more than a corresponding star or delta winding. This is extensively used for earthing transformer.
- Due to **zigzag** connection (interconnection between phases), third harmonic voltages are reduced. It also allows unbalanced loading. The zigzag connection is employed for LV winding. For a given total voltage per phase, the zigzag side requires 15% more turns as compared to normal phase connection. In cases where delta connections are weak due to large number of turns and small cross sections, then zigzag star connection is preferred. It is also used in rectifiers.

<u>Scott connection</u>

There are two main reasons for the need to transform from three phases to two phases,

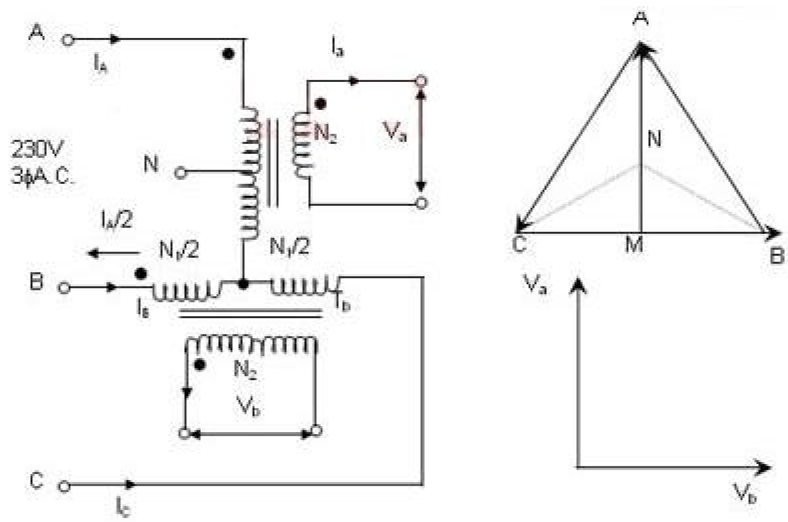
1. To give a supply to an existing two phase system from a three phase supply.

2. To supply two phase furnace transformers from a three phase source.

Two-phase systems can have 3-wire, 4-wire, or 5-wire circuits. It is needed to be considering that a two-phase system is not 2/3 of a three-phase system. Balanced three-wire, two-phase circuits have two phase wires, both carrying approximately the same amount of current, with a neutral wire carrying 1.414 times the currents in the phase wires. The phase-to-neutral voltages are 90° out of phase with each other.

Two phase 4-wire circuits are essentially just two ungrounded single-phase circuits that are electrically 90° out of phase with each other. Two phase 5-wire circuits have four phase wires plus a neutral; the four phase wires are 90° out of phase with each other.

A Scott-T transformer (also called a Scott connection) is a type of circuit used to derive two-phase power from a three-phase source or vice-versa. The Scott connection evenly distributes a balanced load between the phases of the source. Scott T Transformers require a three phase power input and provide two equal single phase outputs called Main and Teaser. The MAIN and Teaser outputs are 90 degrees out of phase. The MAIN and the Teaser outputs must not be connected in parallel or in series as it creates a vector current imbalance on the primary side. MAIN and Teaser outputs are on separate cores. An external jumper is also required to connect the primary side of the MAIN and Teaser sections. The schematic of a typical Scott T Transformer is shown below:



4.23 Connection diagram of Scott-connected transformer and vector relation of input and output

From the phasor diagram it is clear that the secondary voltages are of two phases with equal magnitude and 90° phase displacement.

Scott T Transformer is built with two single phase transformers of equal power rating. Assuming the desired voltage is the same on the two and three phase sides, the Scott-T transformer connection consists of a center-tapped 1:1 ratio main transformer, T1, and an 86.6% (0.5√3) ratio teaser transformer, T2. The center-tapped side of T1 is connected between two of the phases on the three-phase side. Its center tap then connects to one end of the lower turn count side of T2, the other end connects to the remaining phase. The other side of the transformers then connects directly to the two pairs of a two-phase four-wire system.

If the main transformer has a turn's ratio of 1: 1, then the teaser transformer requires a turn's ratio of 0.866: 1 for balanced operation. The principle of operation of the Scott connection can be most easily seen by first applying a current to the teaser secondary windings, and then applying a current to the main secondary winding, calculating the primary currents separately and superimposing the results.

The primary three-phase currents are balanced; i.e., the phase currents have the same magnitude and their phase angles are 120° apart. The apparent power supplied by the main transformer is greater than the apparent power supplied by the teaser transformer. This is easily verified by observing that the primary currents in both transformers have the same magnitude; however, the primary voltage of the teaser transformer is only 86.6% as great as the primary voltage of the main transformer. Therefore, the teaser transforms only 86.6% of the apparent power transformed by the main.

- The total real power delivered to the two phase load is equal to the total real power supplied from the three-phase system, the total apparent power transformed by both transformers is greater than the total apparent power delivered to the two-phase load.
- The apparent power transformed by the teaser is 0.866 X IH1= 1.0 and the apparent power transformed by the main is 1.0X IH2 =1.1547 for a total of 2.1547 of apparent power transformed.
- The additional 0.1547 per unit of apparent power is due to parasitic reactive power owing between the two halves of the primary winding in the main transformer.
- Single-phase transformers used in the Scott connection are specialty items that are virtually impossible to buy ‘‘off the shelf ’’ nowadays. In an emergency, standard distribution transformers can be used.

If desired, a three phase, two phase, or single phase load may be supplied simultaneously using scott-connection. The neutral points can be available for grounding or loading purposes. The Scott T connection in theory would be suitable for supplying a three, two and single phase load simultaneously, but such loads are not found together in modern practice.

The Scott T would not be recommended as a connection for 3 phase to 3 phase applications for the following reasons:

The loads of modern buildings and office buildings are inherently unbalanced and contain equipment that can be sensitive to potential voltage fluctuations that may be caused by the Scott T design.

A properly sized Scott T transformer will have to be a minimum of 7.75% larger than the equivalent Delta-Wye transformer. Properly sized, it would be a bulkier and heavier option and should not be considered a less expensive solution.

Open Delta or V-Connection

As seen previously in connection of three single phase transformers that if one of the transformers is unable to operate then the supply to the load can be continued with the remaining two transformers at the cost of reduced efficiency. The connection that obtained is called V-V connection or open delta connection.

Consider the Fig. 4.24 in which 3 phase supply is connected to the primaries. At the secondary side three equal three phase voltages will be available on no load.

The voltages are shown on phasor diagram. The connection is used when the three phase load is very very small to warrant the installation of full three phase transformer.

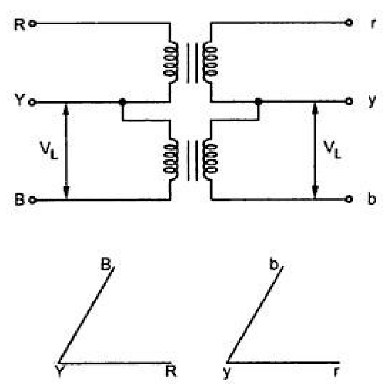


Fig. 4.24 Open delta connection of transformer at noload

If one of the transformers fails in Δ - Δ bank and if it is required to continue the supply even though at reduced capacity until the transformer which is removed from the bank is repaired or a new one is installed then this type of connection is most suitable.

When it is anticipated that in future the load increase, then it requires closing of open delta. In such cases open delta connection is preferred. It can be noted here that the removal of one of the transformers will not give the total load carried by V - V bank as two third of the capacity of Δ - Δ bank.

The load that can be carried by V - V bank is only 57.7% of it.

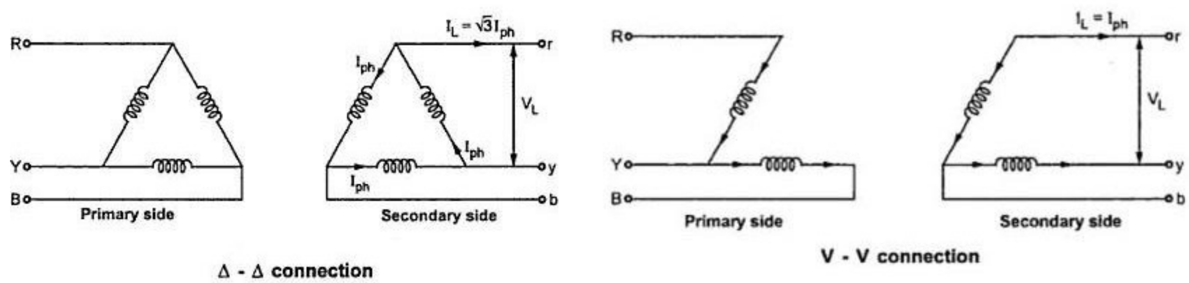


Fig. 4.25 Delta-delta and V-V connection

It can be seen from the Fig. 4.25 of delta delta connection that

$$\Delta - \Delta \text{ capacity} = \sqrt{3}\, V_L I_L = \sqrt{3}\, V_L (\sqrt{3}\, I_{ph})$$
$$\Delta - \Delta \text{ capacity} = 3\, V_L I_{ph}$$

It can also be noted from the Fig. 4.25 V-V connection that the secondary line current I_L is equal to the phase current I_{ph}.

$$\text{V- V capacity} = \sqrt{3}\, V_L I_L = \sqrt{3}\, V_L I_{ph}$$

$$\text{So, } \frac{\text{V-V capacity}}{\Delta - \Delta \text{ capacity}} = \frac{\sqrt{3} V_L I_{ph}}{3 V_L I_{ph}} = \frac{1}{\sqrt{3}} = 0.577 \simeq 58\%$$

Thus the three phase load that can be carried without exceeding the ratings of the transformers is 57.5 percent of the original load. Hence it is not 66.7 % which was expected otherwise.

The reduction in the rating can be calculated as {(66.67 - 57.735)/(57.735)}x 100 = 15.476

Suppose that we consider three transformers connected in Δ - Δ fashion and supplying their rated load. Now one transformer is removed then each of the remaining two transformers will be overloaded. The overload on each transformer will be given as,

$$\frac{\text{Total load in V-V}}{\text{VA rating of each transformer}} = \frac{\sqrt{3} V_L I_{ph}}{V_L I_{ph}} = \sqrt{3} = 1.732$$

This overload can be carried temporarily if provision is made to reduce the load otherwise overheating and breakdown of the remaining two transformers would take place.

- <u>The limitation with V -V connection are given below :</u>

The average p.f. at which V- V bank is operating is less than that with the load . This power p.f is 86.6 % of the balanced load p.f.

- The two transformers in V -V bank operate at different power factor except for balanced unity p.f .load.
- The terminals voltages available on the secondary side become unbalanced. This may happen even though load is perfectly balanced.
- Thus in summary we can say that if tow transformers are connected in V - V fashion and are loaded to rated capacity and one transformer is added to increase the total capacity by √3 or 173.2 %. Thus the increase in capacity is 73.2 % when converting from a V - V system to a Δ-Δ system.
- With a bank of tow single phase transformers connected in V-V fashion supplying a balanced 3 phase load with cosΦ asp.f., one of the transformer operate at a p.f. of cos (30-Φ) and other at cos (30+Φ). The powers of tow transformers are given by,

$$P_1 = KVA \cos(30-\Phi)$$

$$P_2 = KVA \cos(30+\Phi)$$

Oscillating Neutral

In addition to the operation of transformers on the sinusoidal supplies, the harmonic behavior becomes important as the size and rating of the transformer increases. The effects of the harmonic currents are

1. Additional copper losses due to harmonic currents
2. Increased core losses
3. Increased electro-magnetic interference with communication circuits.

On the other hand the harmonic voltages of the transformer cause

1. Increased dielectric stress on insulation
2. Electro static interference with communication circuits.

3. Resonance between winding reactance and feeder capacitance.

In the present times a greater awareness is generated by the problems of harmonic voltages and currents produced by non-linear loads like the power electronic converters. These combine with non-linear nature of transformer core and produce severe distortions in voltages and currents and increase the power loss. Thus the study of harmonics is of great practical significance in the operation of transformers.

In the case of single phase transformers connected to form three phase bank, each transformer is magnetically decoupled from the other. The flow of harmonic currents are decided by the type of the electrical connection used on the primary and secondary sides. Also, there are three fundamental voltages in the present case each displaced from the other by 120 electrical degrees. Because of the symmetry of the a.c. wave about the time axis only odd harmonics need to be considered. The harmonics which are triplen (multiples of three) behave in a similar manner as they are co-phasal or in phase in the three phases. The non-triplen harmonics behave in a similar manner to the fundamental and have ±120° phase displacement between them.

When the connection of the transformer is Yy without neutral wires both primary and secondary connected in star no closed path exists. As the triplen harmonics are always in phase, by virtue of the Y connection they get canceled in the line voltages. Non-triplen harmonics like fundamental, become 0 times phase value and appear in the line voltages. Line currents remain sinusoidal except for non-triplen harmonic currents. Flux wave in each transformer will be flat topped and the phase voltages remain peaked. The potential of the neutral is no longer steady. The star point oscillates due to the third harmonic voltages. This is termed as "oscillating neutral".

Tertiary winding

Apart from the Primary & Secondary windings, there sometimes placed a third winding in power transformers called "Tertiary Winding". Its purpose is to provide a circulating path for the harmonics (especially third harmonics) produced in the transformers along with power frequency (50Hz. third harmonic means 150 Hz oscillations). In delta-delta, delta-star and star-delta transformers all voltages are balanced and there is no floating of neutral or oscillating neutral. The floating of neutral is developed in the case star-star connection only. The transformers are sometimes constructed with three windings. The main windings are connected to form star-star connection and the third winding known as tertiary winding is used to make a closed delta connection to stabilize the neutrals of both primary and secondary circuits. The tertiary winding carries the third-harmonic currents.

Three Winding Transformers

Thus far we have looked at transformers which have one single primary winding and one single secondary winding. But the beauty of transformers is that they allow us to have more than just one winding in either the primary or secondary side. Transformers which have three winding are known commonly as **Three Winding Transformers**.

The principal of operation of a *three winding transformer* is no different from that of an ordinary transformer. Primary and secondary voltages, currents and turns ratios are all calculated the same, the difference this time is that we need to pay special attention to the voltage polarities of each coil winding, the dot convention marking the positive (or negative) polarity of the winding, when we connect them together.

Three winding transformers, also known as a three-coil, or three-winding transformer, contain one primary and two secondary coils on a common laminated core. They can be either a single-phase transformer or a three-phase transformer, (three-winding, three-phase transformer) the operation is the same.

Three Winding Transformers can also be used to provide either a step-up, a step-down, or a combination of both between the various windings. In fact a three winding transformers have two secondary windings on the same core with each one providing a different voltage or current level output.

As transformers operate on the principal of mutual induction, each individual winding of a three winding transformer supports the same number of volts per turn, therefore the volt-ampere product in each winding is the same, that is $N_P/N_S = V_P/V_S$ with any turns ratio between the individual coil windings being relative to the primary supply.

In electronic circuits, one transformer is often used to supply a variety of lower voltage levels for different components in the electronic circuitry. A typical application of three winding transformers is in power supplies and Triac Switching Converters. So a transformer have two secondary windings, each of which is electrically isolated from the others, just as it is electrically isolated from the primary. Then each of the secondary coils will produce a voltage that is proportional to its number of coil turns.

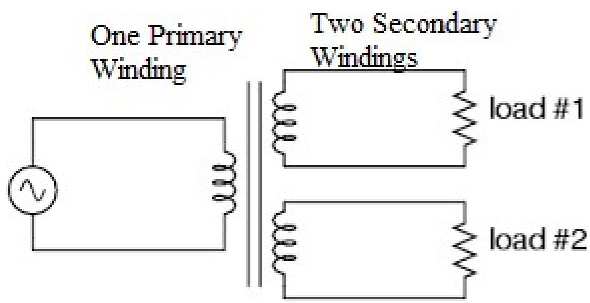


Fig. 4.27 A three winding transformer

The secondary windings can be connected together in various configurations producing a higher voltage or current supply. It must be noted that connecting together transformer windings is only possible if the two windings are electrically identical. That is their current and voltage ratings are the same.

Parallel operation of three phase transformer

4.10.1 Advantages of using transformers in parallel

1. To maximize electrical power system efficiency: Generally electrical power transformer gives the maximum efficiency at full load. If we run numbers of transformers in parallel, we can switch on only those transformers which will give the total demand by running nearer to its full load rating for that time. When load increases, we can switch none by one other transformer connected in parallel to fulfill the total demand. In this way we can run the system

with maximum efficiency.

2. To maximize electrical power system availability: If numbers of transformers run in parallel, we can shut down any one of them for maintenance purpose. Other parallel transformers in system will serve the load without total interruption of power.

3. To maximize power system reliability: If any one of the transformers run in parallel, is tripped due to fault of other parallel transformers is the system will share the load, hence power supply may not be interrupted if the shared loads do not make other transformers over loaded.

4. To maximize electrical power system flexibility: There is always a chance of increasing or decreasing future demand of power system. If it is predicted that power demand will be increased in future, there must be a provision of connecting transformers in system in parallel to fulfill the extra demand because, it is not economical from business point of view to install a bigger rated single transformer by forecasting the increased future demand as it is unnecessary investment of money. Again if future demand is decreased, transformers running in parallel can be removed from system to balance the capital investment and its return.

Conditions for parallel operation

Certain conditions have to be met before two or more transformers are connected in parallel and share a common load satisfactorily. They are,

1. The voltage ratio must be the same.
2. The per unit impedance of each machine on its own base must be the same.
3. The polarity must be the same, so that there is no circulating current between the transformers.
4. The phase sequence must be the same and no phase difference must exist between the voltages of the two transformers.

- **Same voltage ratio :** Generally the turns ratio and voltage ratio are taken to be the same. If the ratio is large there can be considerable error in the voltages even if the turns ratios are the same. When the primaries are connected to same bus bars, if the secondaries do not show the same voltage, paralleling them would result in a circulating current between the secondaries. Reflected circulating current will be there on the primary side also. Thus even without connecting a load considerable current can be drawn by the transformers and they produce copper losses. In two identical transformers with percentage impedance of 5 percent, a no-load voltage difference of one percent will result in a circulating current of 10 percent of full load current. This circulating current gets added to the load current when the load is connected resulting in unequal sharing of the load. In such cases the combined full load of the two transformers can never be met without one transformer getting overloaded.

- **Per unit impedance:** Transformers of different ratings may be required to operate in parallel. If they have to share the total load in proportion to their ratings the larger machine has to draw more current. The voltage drop across each machine has to be the same by virtue of their connection at the input and the output ends. Thus the larger machines have smaller impedance and smaller machines must have larger ohmic impedance. Thus the impedances must be in the inverse ratios of the ratings. As the voltage drops must be the same the per unit impedance of each transformer on its own base, must be equal. In addition if active and reactive power are required to be shared in proportion to the ratings the impedance angles also must be the same. Thus we have the requirement that per unit resistance and per unit reactance of both the transformers must be the same for proper load sharing.
- **Polarity of connection:** The polarity of connection in the case of single phase transformers can be either same or opposite. Inside the loop formed by the two secondaries the resulting voltage must be zero. If wrong polarity is chosen the two voltages get added and short circuit results. In the case of polyphase banks it is possible to have permanent phase error between the phases with substantial circulating current. Such transformer banks must not be connected in parallel. The turns ratios in such groups can be adjusted to give very close voltage ratios but phase errors cannot be compensated. Phase error of 0.6 degree gives rise to one percent difference in voltage. Hence poly phase transformers belonging to the same vector group alone

must be taken for paralleling.

Transformers having —30° angle can be paralleled to that having +30° angle by reversing the phase sequence of both primary and secondary terminals of one of the transformers. This way one can overcome the problem of the phase angle error.

- **Phase sequence-** The phase sequence of operation becomes relevant only in the case of poly phase systems. The poly phase banks belonging to same vector group can be connected in parallel. A transformer with +30° phase angle however can be paralleled with the one with —30° phase angle, the phase sequence is reversed for one of them both at primary and secondary terminals. If the phase sequences are not the same then the two transformers cannot be connected in parallel even if they belong to same vector group. The phase sequence can be found out by the use of a phase sequence indicator.

Load Sharing

When the transformers have equal voltage ratios, the magnitudes of secondary no-load voltages are equal. Further if the primary leakage impedance drops due to exciting currents are also equal, then $\overline{E_a} = \overline{E_b}$ and the circulating current at no load is zero.

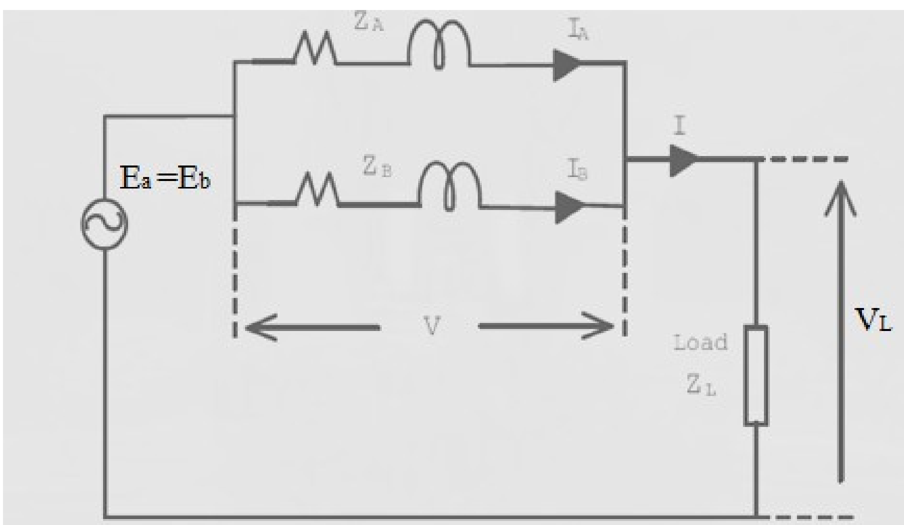


Fig. 4.28 Circuit modelling of two transformer in parallel

The equivalent circuit of two three phase transformer connected in parallel connected with a load of

Z_L impedance on per phase basis is drawn in fig 4.28. In this figure transformer A and B are operating in parallel. I_A and I_B are the load current of the two transformer.

The voltage equation of transformer A is

$$\bar{E}_a - \bar{I}_a\bar{Z}_a = \bar{V}_L = \bar{I}\bar{Z}_L$$

Since $\bar{E}_a = \bar{E}_b$; $\bar{E}_b - \bar{I}_a\bar{Z}_a = \bar{V}_L = \bar{I}\bar{Z}_L$

The voltage equation of transformer B is

$$\bar{E}_b - \bar{I}_b\bar{Z}_b = \bar{V}_L = \bar{I}\bar{Z}_L$$
$$\bar{E}_b - \bar{I}_a\bar{Z}_a = \bar{E}_b - \bar{I}_b\bar{Z}_b$$
$$\bar{I}_a\bar{Z}_a = \bar{I}_b\bar{Z}_b$$

According to the voltage drops across the two equivalent leakage impedance Za and Zb are equal.

According to KCL we can write

$$\bar{I} = \bar{I}_a + \bar{I}_b = \bar{I}_a + \frac{\bar{I}_a\bar{Z}_a}{\bar{Z}_b}$$

$$\bar{I}_a = \bar{I}\frac{\bar{Z}_b}{\bar{Z}_a + \bar{Z}_b}$$

similarly, $\bar{I}_b = \bar{I}\dfrac{\bar{Z}_a}{\bar{Z}_a + \bar{Z}_b}$

Multiplying both the current equations by terminal voltage we get,

$$\bar{S}_a = \bar{S}\frac{\bar{Z}_b}{\bar{Z}_a + \bar{Z}_b}$$

similarly, $\bar{S}_b = \bar{S}\dfrac{\bar{Z}_a}{\bar{Z}_a + \bar{Z}_b}$

Thus the power sharing in between two transformer is given in above equation in VA rating.

Acknowledgement

The committee members gratefully acknowledge google, scribd, NPTEL, openoffice, sumantra pdf, scilab for myriad suggestions and help for preparing this lecture note. The committee members also wants to express their gratitude to the persons out there who thinks knowledge should be free and be accessible and sharable without any restrictions so that every single person on this planet has the same opportunity to explore, expand and become enlightened by the collective gifts of mankind.

However apart from this lecture note students/readers are strongly recommended to follow the below mentioned books in the references and above all confer with the faculty for thorough knowledge of this authoritative subject of electrical engineering.

References

[1] P.S. Bimbhra, Electrical Machinery, Khanna Pub, 7th Ed.

[2] P.S. Bimbhra, Generalized Theory of Electrical Machine, Khanna Pub., 2nd Ed.

[3] M.G. Say, The Performance and Design of Alternating Current Machines, CBS Pub., 3rd Ed.

[4] A.E.Fitzgerald, S.D.Umans, C.Kingsley Jr., Electric Machinery, McGraw Hill Education(india) pvt. Ltd., 6th Ed.

[5] A.S. langsdorf, Theory of Alternating Current Machinery, McGraw Hill Education(india) pvt. Ltd., 2nd Ed.

[6]T.C. Lloyd, A.F. Puchstein and A.G. Conrad, Alternating-Current Machines, John Wiley& Sons pub., 3rd Ed.

[7] Irving Kosow, Electric Machinery and Transformers, Pearson India, 2nd Ed.

[8] C.L.Dawes, A course in Electrical Engineering, Vol-I, Vol-II, Mcgraw Hill Book Company Inc.

[9] P.K. Mukherjee & S. Chakravorty, Electrical Machines, Dhanpat Rai Publications, 2nd Ed.

[10] I.J Nagrath & D.P.Kothari, Electrical Machines, McGraw Hill Education(india) pvt. Ltd., 4th Ed.

[11] N.N.Parker Smith, Problems in Electrical Engineering, CBS Pub. 9th Ed.

<u>Best of Luck to All the Students</u>

GANDHI ACADEMY OF TECHNOLOGY AND ENGINEERING

GOLANTHARA, BERHAMPUR

LECTURE NOTES

ON

ELECTRICAL MACHINE

For 4th Semester

ELECTRONICS AND TELECOMMUNICATION

(As per Syllabus prescribed by SCTE&VT, Odisha)

Prepared By

Mrs. Soumya Shyamali Mahapatra

(Lecturer in Electrical Engineering)

DC Generators

Principle of operation of DC Generator

A D.C generator as shown in figure below the armature be driven by a prime mover in the clock wise direction and the stator field is excited to produce the field poles as shown. There will be induced voltage in each armature conductor. The direction of the induced voltage can be determined by applying *Fleming's right hand rule.* All the conductors under the influence of North Pole will have ⊗ directed induced voltage, while the conductors under the influence of South Pole will have ⊙ induced voltage in them. For a loaded generator the direction of the armature current will be same as that of the induced voltages. Thus ⊗ and ⊙ also represent the direction of the currents in the conductors. We know, a current carrying conductor placed in a magnetic field experiences force, the direction of which can be obtained by applying *Fleming's left hand rule.* Applying this rule to the armature conductors in fig 1.9, the rotor experiences a torque (T_e) in the counter clockwise direction (i.e., opposite to the direction of rotation) known as back torque. For steady speed operation back torque is equal to the machines input torque (T_{pm}) i.e. the torque supplied by prime mover.

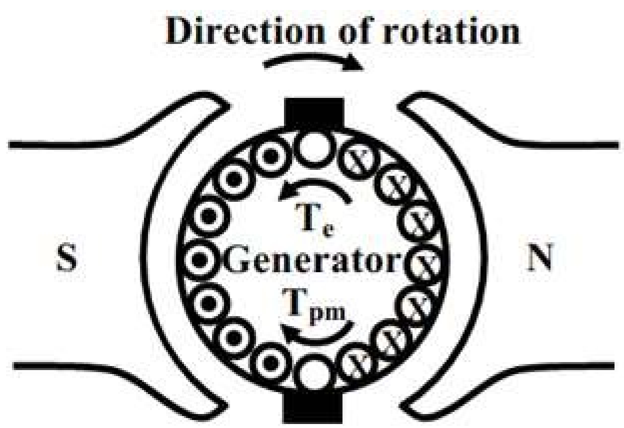


Fig. 1.9 Action of DC generator

Action of Commutator

In DC machines the current in each wire of the armature is actually alternating, and hence a device is required to convert the alternating current generated in the DC generator by electromagnetic induction into direct current, or at the armature of a DC motor to convert the input direct current into alternating current at appropriate times, as illustrated in Fig. 1.10.

DC generator: induced AC *emf* is converted to DC voltage;

DC motor: input direct current is converted to alternating current in the armature at appropriate times to produce a unidirectional torque. The commutator consists of insulated copper segments mounted on an insulated tube. Armature coils are connected in series through the commutator segments. Two brushes are pressed to the commutator to permit current flow. The brushes are placed in the neutral zone, where the magnetic field is close to zero, to reduce arcing. The *commutator* switches the current from one rotor coil to the adjacent coil. The switching requires the interruption of the coil current. The sudden interruption of an inductive current generates high voltages. The high voltage produces flashover and arcing between the commutator segment and the brush.

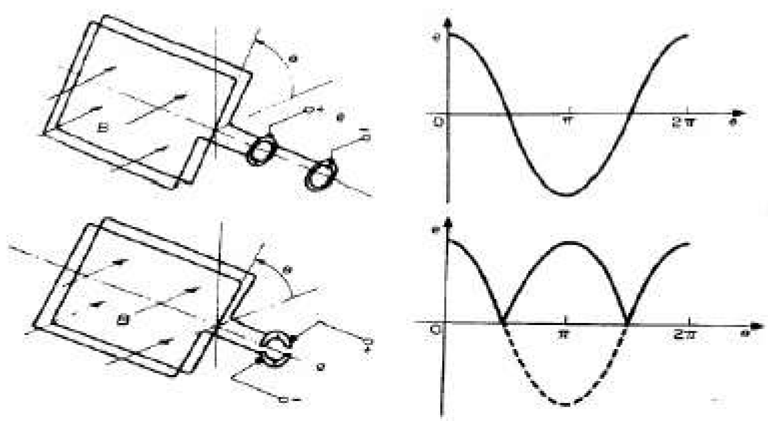


Fig. 1.10Action of Commutator

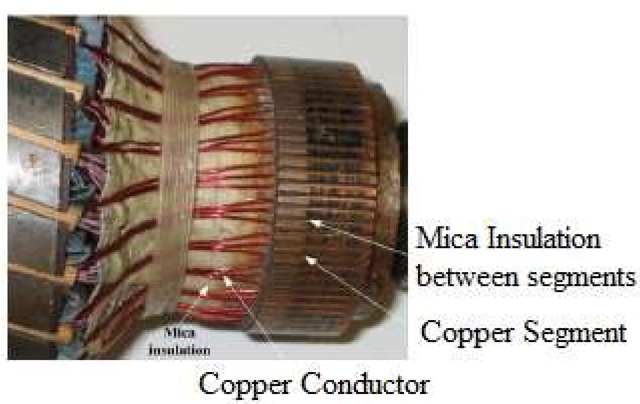


Fig. 1.11 Mechanical view of commutator

Constructional Features

The stator of the dc machine has poles, which are excited by dc current to produce magnetic fields. In the neutral zone, in the middle between the poles, commutating poles or interpoles are placed to reduce sparking of the commutator due to armature reaction. The commutating poles are supplied by dc current. Compensating windings are mounted on the main poles. Field poles are mounted on an iron core that provides a closed magnetic circuit. The motor housing supports the iron core, the brushes and the bearings. The rotor has a ring-shaped laminated iron core with slots. Coils with several turns are placed in the slots. The distance between the two legs of the coil is about 180 electric degrees for full pitch. The coils are connected in series through the commutator segments. The ends of each coil are connected to a commutator segment.

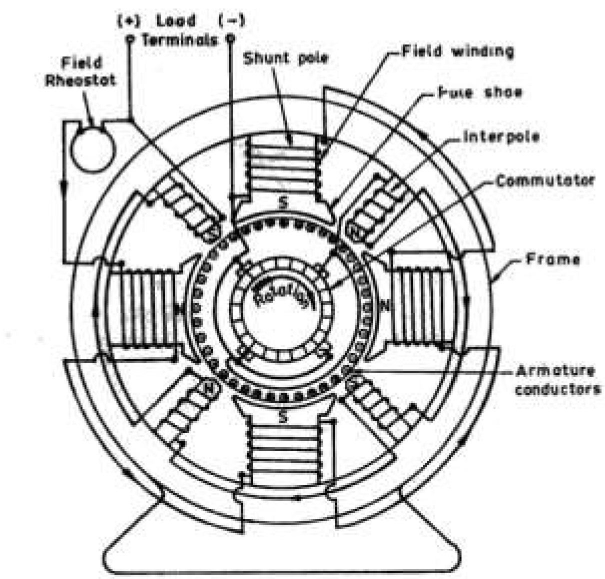


Fig. 1.12 Cross sectional view of DC Machine

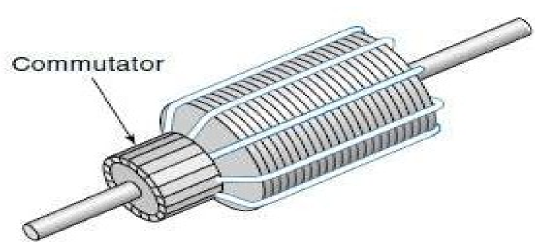


Fig. 1.13 Armature of DC Machine

Armature Winding

DC machines armature consists of armature conductors. The conductors distributed in slots provided on the periphery of the armature is called armature winding. Depending on the way in which the coils are interconnected at the commutator end of the armature, the windings can be classified as lap and wave windings. Further they can be classified as simplex and multiplex.

Coil Span/Coil Pitch:

It represents the span of the coil. For full pitched winding, the span is 180^0 electrical or number of slots per pole. Coil pitch can be represented in terms of electrical degrees, slots or conductor. A full pitched coil leads to maximum voltage per coil.

Back Pitch *(Yb)*:

It is the distance measured in between the two coil sides of the same coil at the back end of the armature, the commutator end being the front end of armature. It can be represented in terms of number of slots or coil sides. Back pitch also represents the span of coil.

Front Pitch (*Yf*):

The distance between the two coil sides of two different coils connected in series at the front end of the armature is called front pitch.

Lap Winding

Lap winding is suitable for low voltage high current machines because of more number of parallel paths. The number of parallel path in lap winding is equal to number of poles.

$$A=P$$

Equalizing rings are connected in lap winding.

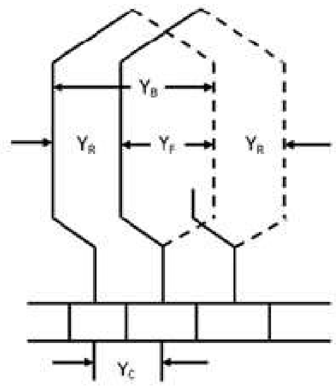


Fig. 1.14 Winding diagram of lap winding

Wave Winding

Wave winding is used for high voltage low current machines. In case of wave winding, the number of parallel path (A) = 2 irrespective of number of poles. Each path will have conductors connected in series.

Equalizing rings are not required in wave winding.

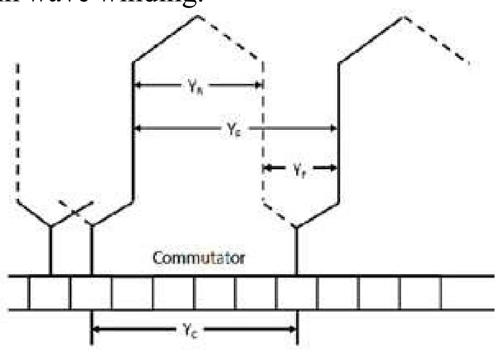


Fig. 1.15 Winding diagram of wave winding

Simplex and Multiplex Winding

Fig 1.14 and Fig. 1.15 shows simplex lap and simplex wave winding.

The degree of multiplicity of a multiplex winding indicates the relative number of parallel paths with respect to the number of parallel paths in the corresponding simplex winding. For example a duplex lap or wave winding is a lap or wave winding having twice as many as parallel paths as a simplex lap or wave winding respectively. The winding can be triplex or quadruplex winding in similar manner.

Use of Laminated Armature

The armature winding of DC machine should be laminated to reduce eddy current losses. The armature body (rotor) rotates in the field magnetic field. Thus in the core of the armature voltage induced which in turn causes current to flow in the body. This current is known as eddy current. This current causes loss and thus heat will be generated. This loss depends on the amount of current flow. To reduce the amount of current flow the resistance of the body should be increased. Thus using lamination the resistance of the path through which current flows will be increased. The amount of eddy current will be reduced and thus eddy current loss can be minimized.

EMF Equation

Let ϕ = flux per pole in weber

Z = number of armature conductors = Number of slots X conductors per slot.

P = Number of poles; A= Number of parallel paths in armature.

A= P for lap wound armature; A=2 for wave wound armature

N = speed of armature in rpm; E = induced emf in each parallel path.

Average emf generated/conductor in one revolution = $\frac{d\phi}{dt}$

Flux cut by a conductor in one revolution = $d\phi = P\phi$ weber.

Since Number of revolutions/second= $\frac{N}{60}$

Time taken for one revolution = $dt = \frac{60}{N}$ seconds

EMF generated/conductor = $\frac{d\phi}{dt} = \frac{P\phi}{\frac{60}{N}} = \frac{P\phi N}{60}$

Since each path has $\frac{Z}{A}$ conductors in series,

EMF generated in each path is $E = \frac{P\phi N}{60} \times \frac{Z}{A}$

$$\boxed{E = \frac{P\phi ZN}{60A}}$$

Methods of excitation

DC machines are excited in two ways-

Separate excitation:

When the field winding is connected to an external source to produce field flux. According to the type of excitation this machines are called separately excited dc machine.

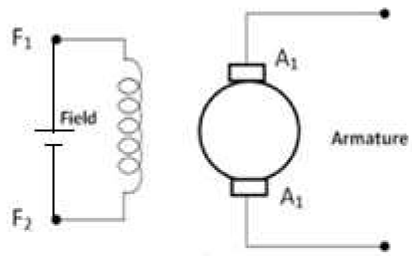


Fig. 1.16 Schematic diagram of separately excited dc machine

Self-excitation:

When the field winding is connected with the armature to produce field flux. A self-excited machine requires residual magnetism for operation. According to the type of excitation this machines are called self-excited dc machine.

Depending on the type of field winding connection DC machines can be classified as:

Shunt machine:

The field winding consisting of large number of turns of thin wire is usually excited in parallel with armature circuit and hence the name shunt field winding. This winding will be having more resistance and hence carries less current.

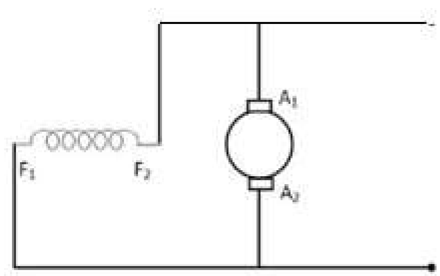


Fig. 1.17 Schematic diagram of dc shunt machine

Series machine:

The field winding has a few turns of thick wire and is connected in series with armature.

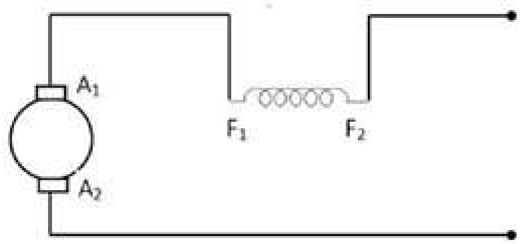


Fig.1.18 Schematic diagram of dc series machine

Compound machine:

Compound wound machine comprises of both series and shunt windings and can be either short shunt or long shunt, cumulative, differential or flat compounded.

Fig. 1.19 A A Schematic diagram of short -shunt compound machine

Fig.1.19 B Schematic diagram of long-shunt compound machine

Build-up of E.M.F

When the armature is rotating with armature open circuited, an emf is induced in the armature because of the residual flux. When the field winding is connected with the armature, a current flows through the field winding (in case of shunt field winding, field current flows even on No-load and in case of series field winding only with load) and produces additional flux. This additional flux along with the residual flux generates higher voltage. This higher voltage circulates more current to generate further higher voltage. This is a cumulative process till the saturation is attained.

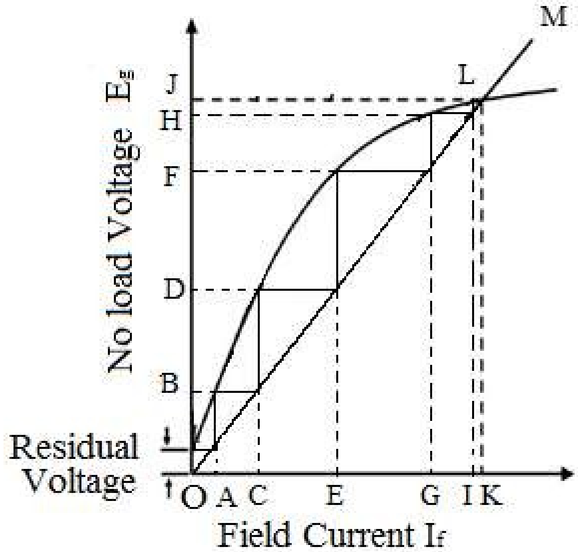


Fig. 1.20 Process of voltage build-up in DC generator

Here OM is the field resistance curve in Fig. 1.20. Initially there will be residual voltage which will create OA field current. This field current will increase the existing magnetic field and the induced voltage will increase up-to OB. This OB voltage will further applied to the field winding and increase the field current to OC. This process will continue upto the point L where the emf curve intersect with field resistance and finally the induced voltage will be OJ. This way voltage builds-up in dc generator.

1.19 Critical Resistance:-

The voltage to which it builds is decided by the resistance of the field winding as shown in the figure 1.21. If field circuit resistance is increased such that the resistance line does not cut OCC like 'OP' in the figure1.21, then the machine will fail to build up voltage to the rated value. The slope of the air gap line drawn as a tangent (OQ) to the initial linear portion of the curve represents the maximum resistance that the field circuit can have beyond which the machine fails to build up voltage. This value of field circuit resistance is called critical field resistance. The field circuit is generally designed to have a resistance value less than this so that the machine builds up the voltage to the rated value.

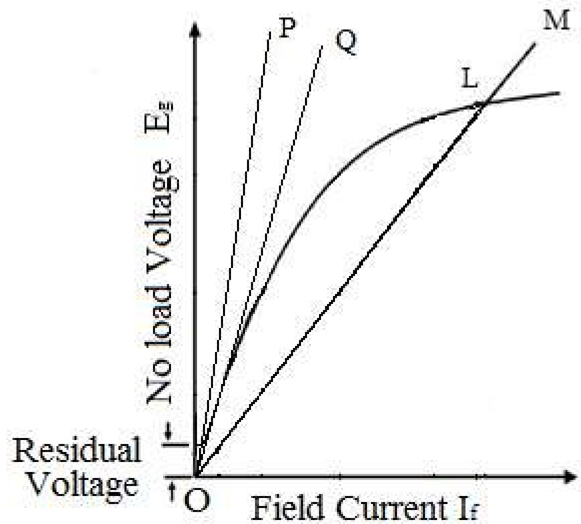
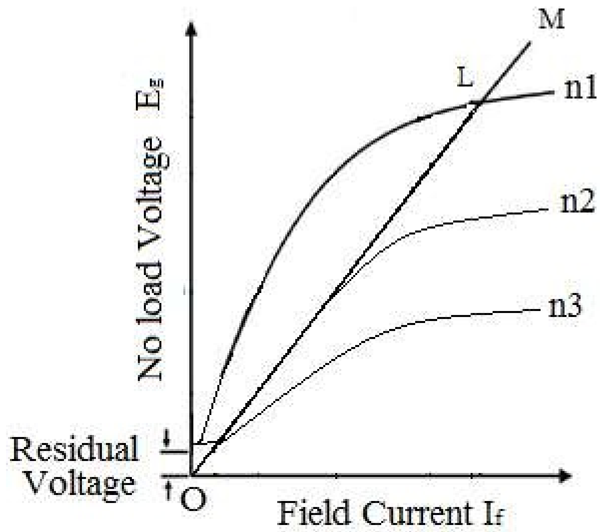


Fig. 1.21 Field current vs No-load voltage for different field resistances

Critical field resistance is defined as the maximum field circuit resistance for a given speed with which the shunt generator would excite. The shunt generator will build up voltage only if field circuit resistance is less than critical field resistance.

Critical Speed:

Voltage of a dc generator is proportional to its speed. Thus when speed will be reduced then the induced voltage will reduced. There can be such situation occur when the speed will be so low that the existing field winding resistance voltage bulid up will not occur. The speed of the generator can be lowered upto a certain level. This minimum value of the speed of the generator for which the generator can excite is called critical speed. It can also define as that speed of a generator for which the existing field resistance of generator becomes its critical field resistance.



connected to the
dual flux.

Fig. 1.22 Field current vs No-load voltage for different speed

In the above figure it is showing that when speed of the generator changes from n1 to n2 and then n3 emf production changes accordingly. Here n1>n2>n3. For speed n3 voltage build-up is not possible. The speed n2 is the critical speed. As shown in the figure at speed n2 generator field resistance become its critical field resistance.

Causes for failure to self-excite and its remedial

i. The field poles may not have residual magnetism. Then the generator will fail to excite.

   Then to restore residual magnetism field winding should be connected to an external dc voltage source. This is called flashing of field.

ii. When the direction of rotation is not proper such that flux produced by the field current reinforces the residual magnetism.

   The rotation of the machine has to be reversed.

iii. The field winding resistance is more than critical resistance then the machine will fail to excite.

   The field winding resistance should be less than critical field resistance.

iv. When the speed of the machine is less than critical speed.

   The machine's speed should be more than critical speed.

v. If the field winding connections are such that newly generated field flux is working in opposite to the existing residual magnetism. Then the generator will fail to excite.

Then the field winding connection should be reversed.

Armature Reaction

The action of magnetic field set up by armature current on the distribution of flux under main poles of a DC machine is called armature reaction.

When the armature of a DC machines carries current, the distributed armature winding produces its own mmf. The machine air gap is now acted upon by the resultant mmf distribution caused by the interaction of field ampere turns (AT_f) and armature ampere turns (AT_a). As a result the air gap flux density gets distorted.

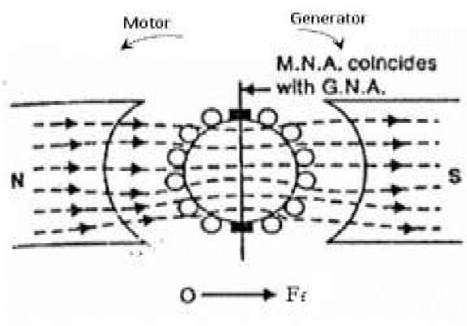


Fig. a

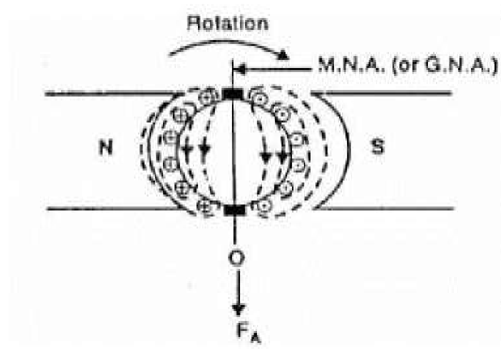


Fig. b

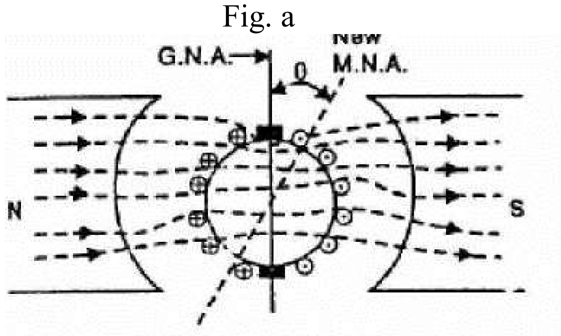


Fig. c

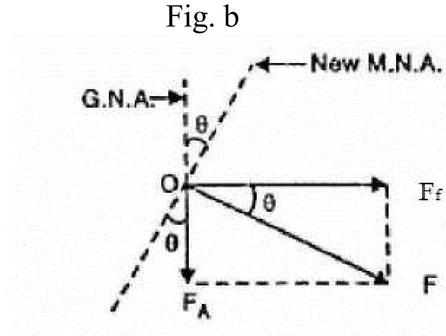


Fig. d

Figure a shows a two pole machine with single equivalent conductor in each slot and the main field mmf (F_f) acting alone. The axis of the main poles is called the direct axis (d-axis) and the interpolar axis is called quadrature axis (q-axis). It can be seen from the Figure b that armature mmf (F_a) is along the interpolar axis. F_a which is at 90^0 to the main field axis is known as cross magnetizing mmf.

Figure c shows the practical condition in which a DC machine operates when both the Field flux and armature flux are existing. Because of both fluxes are acting simultaneously, there is a shift in brush axis and crowding of flux lines at the trailing pole tip and flux lines are weakened or thinned at the leading pole tip. (The pole tip which is first met in the direction of rotation by the armature conductor is leading pole tip and the other is trailing pole tip).

If the iron in the magnetic circuit is assumed unsaturated, the net flux/pole remains unaffected by the armature reaction though the air gap flux density distribution gets distorted. If the main pole excitation is such that the iron is in the saturated region of magnetization (practical case) the increase in flux density at one end of the poles caused by armature reaction is less than the decrease at the other end, so that there is a net reduction in the flux/pole. This is called the demagnetizing effect. Thus it can be summarized that the nature of armature reaction in a DC machine is

1. Cross magnetizing with its axis along the q-axis.
2. It causes no change in flux/pole if the iron is unsaturated but causes reduction in flux/pole in the presence of iron saturation. This is termed as demagnetizing effect. The resultant mmf 'F' is shown in figure d.

1.22.1Cross Magnetizing Ampere Turns/pole(AT_c)

If the brush is shifted by an angle θ as shown in figure 1.23 then the conductors lying in between the angles BOC and DOA are carrying the current in such a way that the direction of the flux is downwards i.e., at right angles to the main flux. This results is the distortion in the main flux. Hence, these conductors are called cross magnetizing or distorting ampere conductors.

Total armature conductors/pole= $\frac{Z}{P}$

Demagnetizing conductors / pole = $Z\frac{2\theta}{360}$

Therefore cross magnetizing conductors/pole= $\frac{Z}{P} - Z\,\frac{2\theta}{360}$

Cross magnetizing ampere turns/pole AT$_c$= $\frac{ZI_a}{a}\left(\frac{1}{2P} - \frac{\theta}{360}\right)$

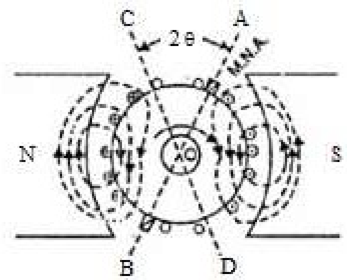


Fig. 1.23 Cross-magnetizing ampere conductors

1.22.2 Demagnetizing Ampere Turns /pole (AT$_d$)

The exact conductors which produce demagnetizing effect are shown in Fig 1.24, Where the brush axis is given a forward lead of θ so as to lie along the new axis of M.N.A. The flux produced by the current carrying conductors lying in between the angles AOC and BOD is such that, it opposes the main flux and hence they are called as demagnetizing armature conductors.

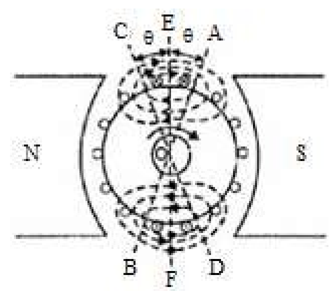


Fig. 1.24 Demagnetizing ampere conductors

Z= total no of armature conductors

Current in each armature conductors= $\frac{I_a}{a}$

θ =Forward lead in mechanical or angular deg.

Total no of armature conductors in between angles AOC & BOD =$\frac{4\theta}{360} Z$

Demagnetizing amp turns/poles AT$_d$ = $\frac{Z\theta I_a}{360a}$

Compensating Winding

Due to armature reaction flux density wave get distorted and reduced. Due to distortion of flux wave the peak flux density increases to such a high value that it creates high induced emf. If this emf is higher than the breakdown voltage across adjacent segments, a spark over could result which can easily spread over the whole commutator, and there will be a ring of fire, resulting in the complete short circuit of the armature.

To protect armature from such adverse condition armature reaction must be neutralized. To neutralize the armature reaction ampere-turns by compensating winding placed in the slots cut out in pole face such that the axis of the winding coincides with the brush axis as shown figure 1.25.

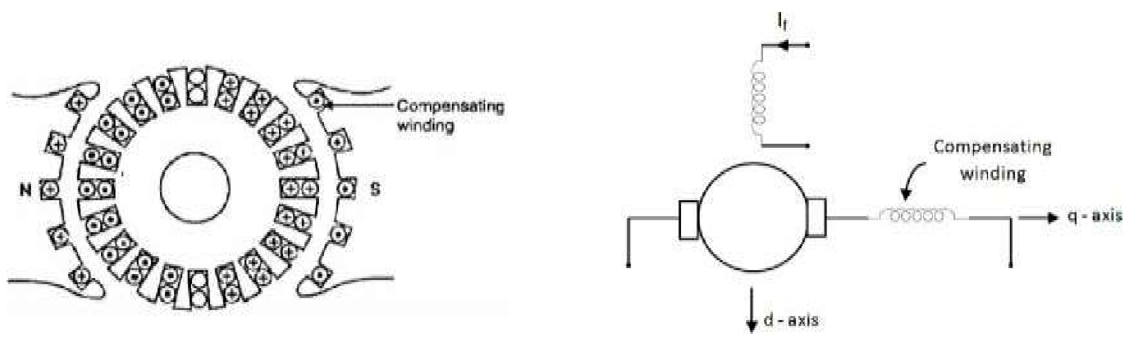


Fig. 1.25 Compensating conductors in field poles and the connection of compensating conductors with armature

The compensating windings neutralize the armature mmf directly under the pole which is the major portion because in the interpole region the air gap will be large. Compensating windings are connected in series with armature so that it will create mmf proportional to armature mmf.

The number of ampere-turns required in the compensating windings is given by

$$AT_c = \text{Total Armature Ampere Turns} \times \frac{\text{Pole Arc}}{\text{Pole Pitch}}$$

Commutation

The process of reversal of current in the short circuited armature coil is called 'Commutation'. This process of reversal takes place when coil is passing through the interpolar axis (q-axis), the coil is short circuited through commutator segments and brush.

The process of commutation of coil 'CD' is shown Fig. 1.26. In sub figure 'c' coil 'CD' carries 20A current from left to right and is about to be short circuited in figure 'd' brush has moved by a small width and the brush current supplied by the coil are as shown. In figure 'e' coil 'CD' carries no current as the brush is at the middle of the short circuit period and the brush current in supplied by coil 'AB' and coil 'EF'. In sub figure 'f' the coil 'CD' which was carrying current from left to right carries current from right to left. In sub fig 'g' spark is shown which is due to the reactance voltage. As the coil is embedded in the armature slots, which has high permeability, the coil possess appreciable amount of self inductance. The current is changed from +20 to –20. So due to self inductance and variation in the current from +20 to –20, a voltage is induced in the coil which is given by L dI/dt. This emf opposes the change in current in coil 'CD' thus sparking occurs.

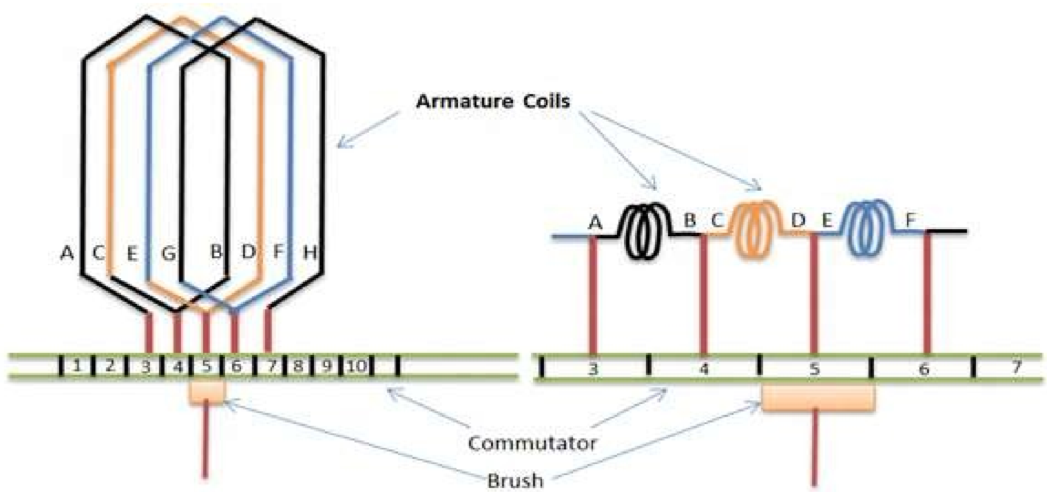


Fig a

Fig b

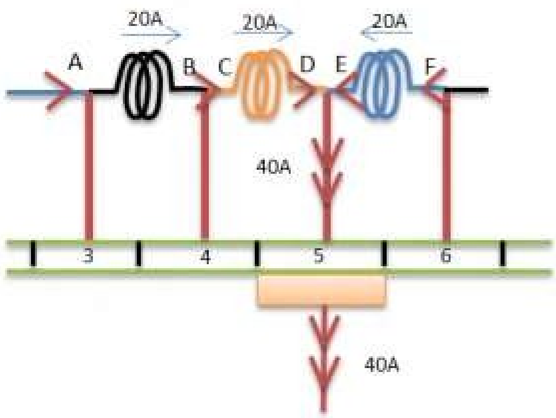


Fig.c

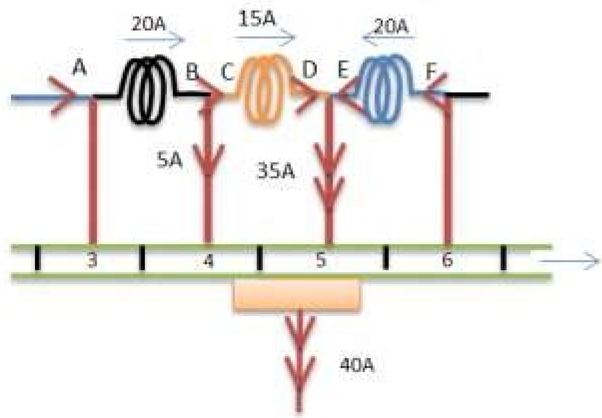


Fig.d

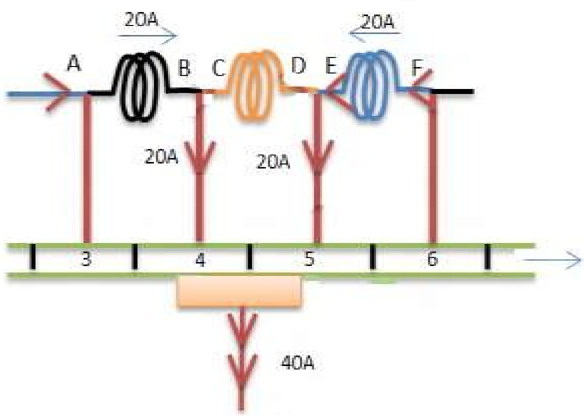


Fig.e

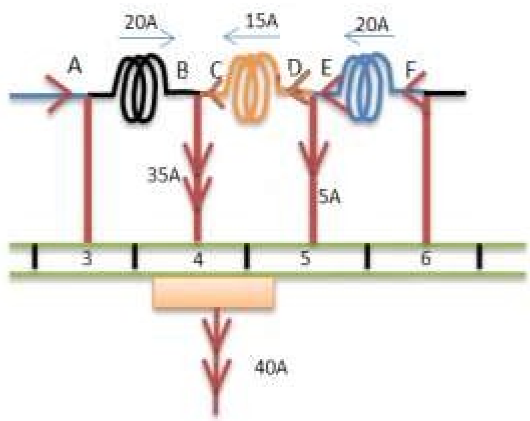


Fig.f

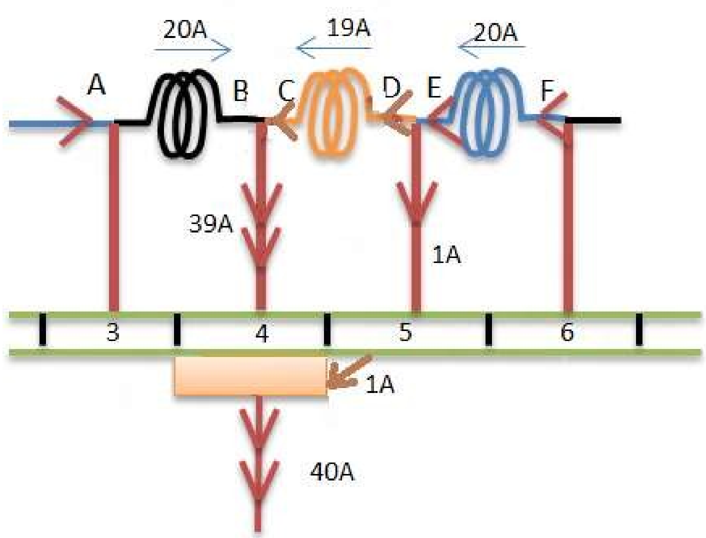


Fig.g

Fig.1.26 a-g shows the process of commutation

Reactance Voltage

During commutation sparking occurs in the commutator segment and brush due to presence of reactance voltage. This voltage is generated due to change of current in the commutating coil for its self-inductance and also due to mutual inductance of the adjacent coils. This voltage is called reactance voltage and according to Lenz's law this induced voltage oppose its cause of production. Here the cause of production is the change in current in the coil under commutation. Thus the commutation becomes poorer.

Reactance voltage = co-efficient of self-inductance X rate of change of current= $L\dfrac{di}{dt}$

Time of short circuit = T_c = (time required by commutator to move a distance equal to the circumferential thickness of brush)–(one mica insulating strip) = Time of commutation

Let Wb= brush width in cm

W_i = width of mica insulation in cm

V_c = peripheral velocity of commutator segments in cm/sec.

Then $T_c = \dfrac{W_b - W_i}{V_c}$ sec

Total change in current = I - (-I) = 2I

Therefore self-induced or reactance voltage = $L\dfrac{2I}{T_c}$ for linear commutation

$=1.11L\dfrac{2I}{T_c}$ for sinusoidal commutation

If brush width is given in terms of commutator segments, then commutator velocity should be converted in terms of commutator segments/seconds.

Method of Improving Commutation

Commutation can be improved in two ways by (i) Resistance commutation

(ii) E.M.F commutation.

Resistance Commutation

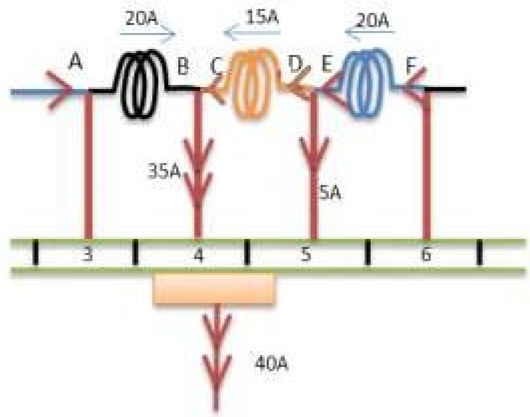


Fig. 1.27

In this method the resistance of the brushes are increased by changing then from copper brush to carbon brush. From the above figure 1.27 it is seen that when current '20A' from coil 'EF' reaches the commutator segment '5', it has two parallel paths opened to it. The first path is straight from bar '5' to the brush and the other is via short circuited coil 'CD' to bar '4' and then to brush. If copper brushes are used the current will follow the first path because of its low contact resistance. But when carbon brushes having high resistance are used, then current '20A' will prefer the second path because the resistance r_1 of first path will increase due to reducing area of contact with bar '5' and the resistance r_2 of second path decreases due to increasing area of contact with bar '4'. Hence carbon brushes help in obtaining sparkles commutation. Also, carbon brushes lubricate and polish commutator. But, because of high resistance the brush contact drop increases and the commutator has to be made larger to dissipate the heat due to loss. Carbon brushes require larger brush holders because of lower current density.

E.M.F commutation:

To neutralize sparking caused by reactance voltage in this method an emf is producied whichacts in opposite direction to that of reactance voltage, so that the reactance voltage is completely eliminated. The neutralization of emf may be done in two ways (i) by giving brush a forward lead sufficient enough to bring the short circuited coil under the influence of next pole of opposite polarity or (ii) by using interpoles or compoles. The second method is commonly employed.

Interpoles or Compoles

These are small poles fixed to the yoke and placed in between the main poles as shown in figure 1.28. They are wound with few turns of heavy gauge copper wire and are connected in series with the armature so that they carry full armature current. Their polarity in case of generator is that of the main pole ahead in the direction of rotation. Their polarity in case of motor is that of the main pole behind in the direction of rotation.

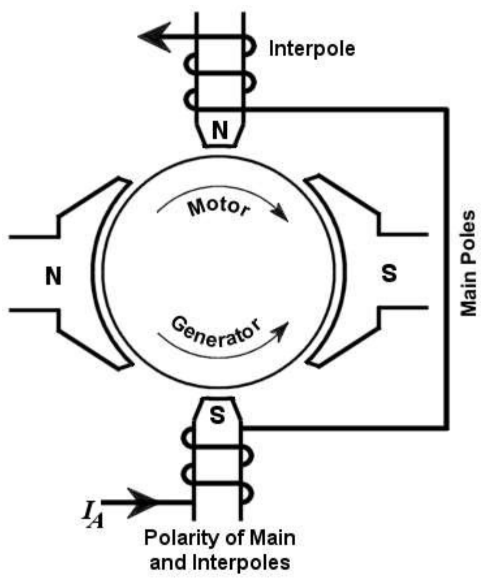


Fig. 1.28 Inter-poles of DC machines

The function of interpoles is (i) to induce an emf which is equal and opposite to that of the reactance voltage. Interpoles neutralize the cross magnetizing effect of armature reaction.

DC Machines Characteristics

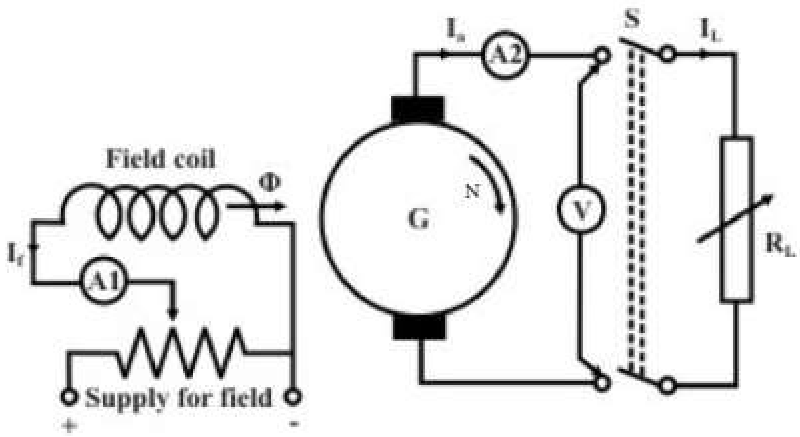
There main characteristics of dc machines are

i. No-load characteristics or Open Circuit Characteristics

ii. Load characteristics

No-load characteristics or Open Circuit Characteristics of separately excited generator

In this type of generator field winding is excited from a separate source, hence field current is independent of armature terminal voltage as shown on figure 1.29. The generator is driven by a prime mover at rated speed, say N rpm. With switch *S* in opened condition, field is excited via a potential divider connection from a separate d.c source and field current is gradually increased. The field current will establish the flux per pole ϕ. The voltmeter V connected across the armature terminals of the machine will record the generated emf ($E_g = \frac{P\phi ZN}{60A} = k\phi N$). As field current is increased, E_g will increase. E_g versus I_f plot at speed N_1, N_2, N_3 is shown in figure below. Where $N_1 > N_2 > N_3$.



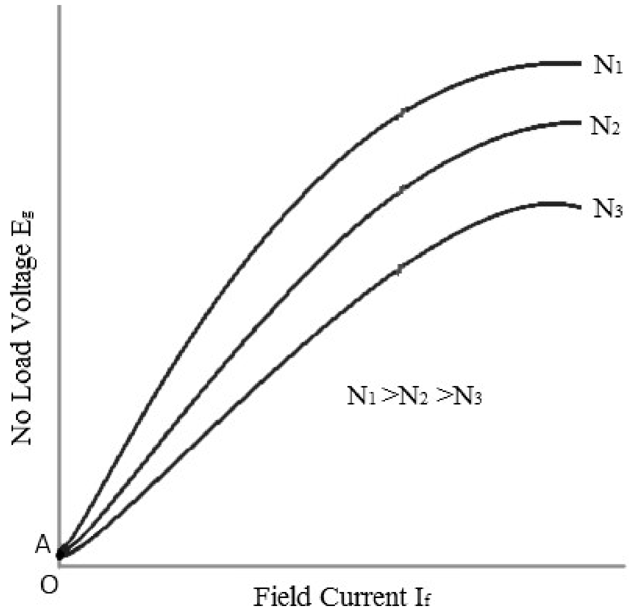


Fig. 1.29 Open circuit characteristics of DC separately excited generator

Load characteristic of separately excited generator

Load characteristic is the characteristics in between terminal voltage V_T with load current I_L of a generator with constant speed and constant field current. For $I_L = 0$, $V_T = E_g$ should be the first point on the load characteristic. With increase of load current the terminal voltage will drop due to armature resistance and reaction drop. In the figure below the rated load current shown by point A. Hence the load characteristic will be drooping in nature as shown in figure 1.30.

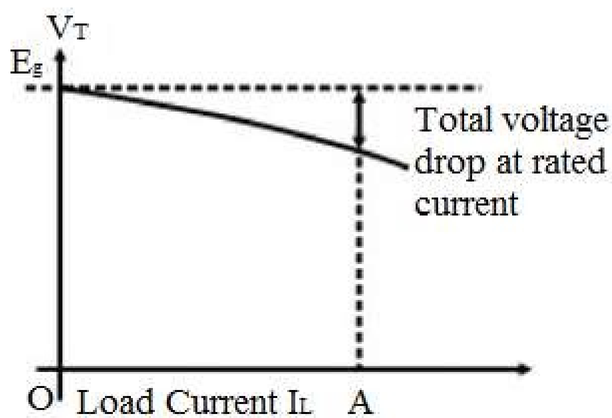


Fig 1.30 Load characteristics of DC separately excited generator

Characteristics of a shunt generator

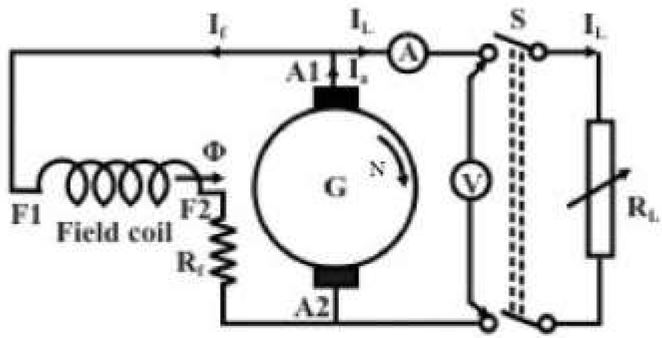


Fig 1.31 Connection diagram to obtain no-load and load characteristic

To obtain OCC of DC shunt generator the above circuit will be used with switch S kept open in fig. 1.31. As this machine is self-excited thus there is no need to use separate dc source for producing field current. The voltage build up process is described earlier. The open nature of the OCC will be similar to separately excited machine.

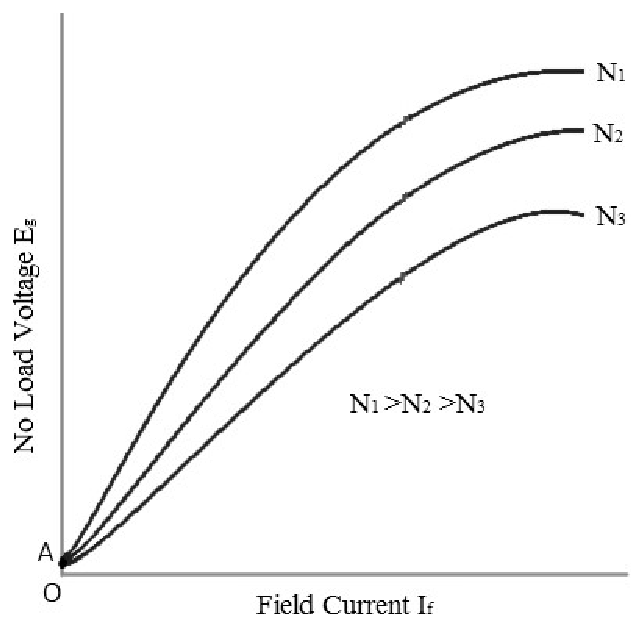


Fig. 1.32 Open circuit characteristics of DC shunt generator

Load characteristic of shunt generator

With switch S in open condition in fig. 1.31, the generator is practically under no load condition as field current is pretty small. The voltmeter reading will be E_g as shown in figures below. In other words, $V_T = E_g$, $I_L = 0$ is the first point in the load characteristic. To load the machine S is closed and the load resistances decreased so that it delivers load current I_L. Unlike separately excited motor, here $I_L \neq I_a$. In fact, for shunt generator, $I_a = I_L - I_f$. So increase of I_L will mean increase of I_a as well. The drop in the terminal voltage will be caused by the usual $I_a r_a$ drop, brush voltage drop and armature reaction effect. Apart from these, in shunt generator, as terminal voltage decreases, field current hence ϕ also decreases causing additional drop in terminal voltage. Remember in shunt generator, field current is decided by the terminal voltage by virtue of its parallel connection with the armature. Figure 1.33 gives the plot of terminal voltage versus load current which is called the load characteristic.

As the load resistance is decreased (load current increased), the terminal voltage drops until point B is reached. If load resistance is further decreased, the load current increases momentarily. This momentary increase in load current produces more armature reaction thus causing a reduction in the terminal voltage and field current. The net reduction in terminal voltage is so large that the load current decreases and the characteristic turns back. In case the machine is short circuited, the curve terminates at point H. Here OH is the load current due to the voltage generated by residual flux.

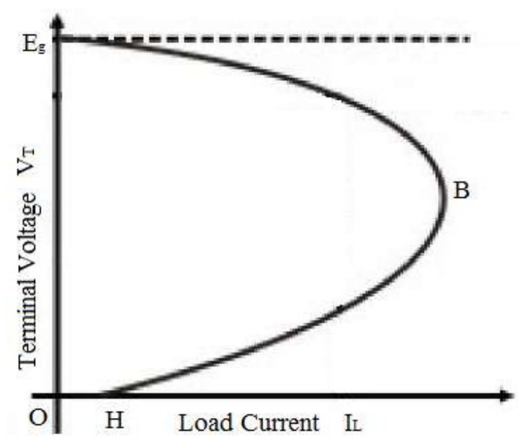


Fig. 1.33 Load characteristics of DC shunt generator

Compound generator

As introduced earlier, compound machines have both series and shunt field coils. Series field coil may be connected in such a way that the mmf produced by it aids the shunt field mmf-then the machine is said to be cumulative compound machine, otherwise if the series field mmf acts in opposition with the shunt field mmf – then the machine is said to be differential compound machine.

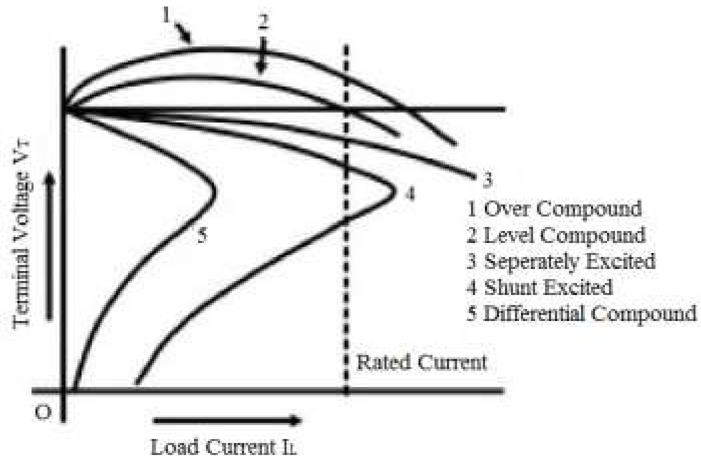


Fig. 1.34 Load characteristics of DC compound generator

In a compound generator, series field coil current is load dependent. Therefore, for a cumulatively compound generator, with the increase of load, flux per pole increases. This in turn increases the generated emf and terminal voltage. Unlike a shunt motor, depending on the strength of the series field mmf, terminal voltage at full load current may be same or more than the no load voltage. When the terminal voltage at rated current is same that at no load condition, then it is called a level compound generator. If however, terminal voltage at rated current is more than the voltage at no load, it is called a over compound generator. The load characteristic of a cumulative compound generator will naturally be above the load characteristic of a shunt generator as depicted in figure 1.34. At load current higher than the rated current, terminal voltage starts decreasing due to saturation, armature reaction effect and more drop in armature and series field resistances.

Parallel Operation of DC Generator

Advantages of DC generator operating in parallel

In a dc power plant, power is usually supplied from several generators of small ratings connected in parallel instead of from one large generator. This is due to the following reasons:

a. ***Continuity of service:***

If a single large generator is used in the power plant, then in case of its breakdown, the whole plant will be shut down. However, if power is supplied from a number of small units operating in parallel, then in case of failure of one unit, the continuity of supply can be maintained by other healthy units.

b. *Efficiency:*

Generators run most efficiently when loaded to their rated capacity. Therefore, when load demand on power plant decreases, one or more generators can be shut down and the remaining units can be efficiently loaded.

c. *Maintenance and repair:*

Generators generally require routine-maintenance and repair. Therefore, if generators are operated in parallel, the routine or emergency operations can be performed by isolating the affected generator while load is being supplied by other units. This leads to both safety and economy.

d. *Increasing plant capacity:*

In the modern world of increasing population, the use of electricity is continuously increasing. When added capacity is required, the new unit can be simply paralleled with the old units.

e. *Non-availability of single large unit:*

In many situations, a single unit of desired large capacity may not be available. In that case a number of smaller units can be operated in parallel to meet the load requirement. Generally a single large unit is more expensive.

Connecting Shunt Generators in Parallel:

The generators in a power plant are connected in parallel through bus-bars. The bus-bars are heavy thick copper bars and they act as +ve and -ve terminals. The positive terminals of the generators are connected to the +ve side of bus-bars and negative terminals to the negative side of bus-bars. Fig. 1.35 shown shunt generator 1 connected to the bus-bars and supplying load. When the load on the power plant increases beyond the capacity of this generator, the second shunt generator 2 is connected in parallel with the first to meet the increased load demand.

The procedure for paralleling generator 2 with generator 1 is as under:

i. The prime mover of generator 2 is brought up to the rated speed. Now switch S2 in the field circuit of the generator 2 is closed.

ii. Next circuit breaker CB-2 is closed and the excitation of generator 2 is adjusted till it generates voltage equal to the bus-bars voltage. This is indicated by voltmeter V2.

iii. Now the generator 2 is ready to be paralleled with generator 1. The main switch DPST2 is closed, thus putting generator 2 in parallel with generator 1. Note that generator 2 is not supplying any load because its generated emf is equal to bus-bars voltage. The generator is said to be “floating” (i.e. not supplying any load) on the bus-bars.

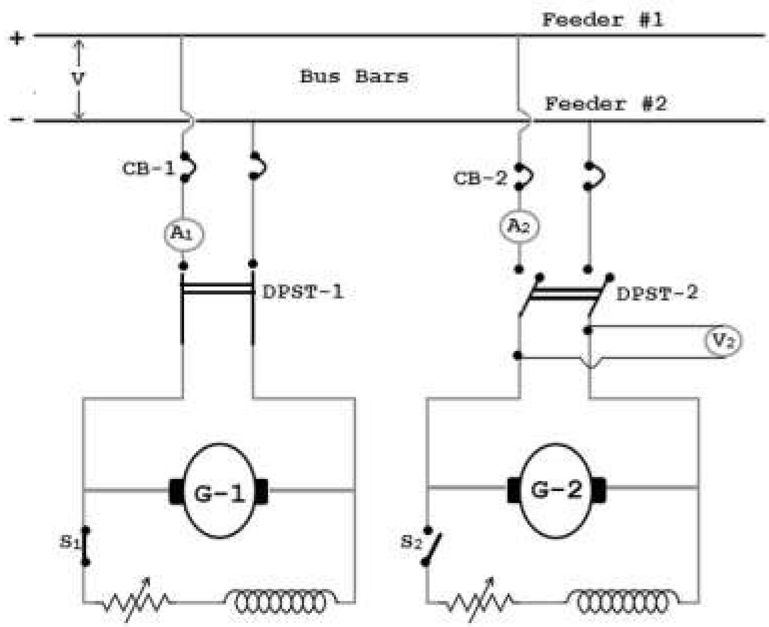


Fig. 1.35 Schematic diagram of DC generator connected in parallel

iv. If generator 2 is to deliver any current, then its generated voltage E_g should be greater than the bus-bars voltage V_T. In that case, current supplied by it is $I = (E_g - V_T)/R_a$ where R_a is the resistance of the armature circuit. By increasing the field current (and hence induced emf E_g), the generator 2 can be made to supply proper amount of load.

v. The load may be shifted from one shunt generator to another merely by adjusting the field excitation. Thus if generator 1 is to be shut down, the whole load can be shifted onto generator 2 provided it has the capacity to supply that load. In that case, reduce the current supplied by generator 1 to zero (This will be indicated by ammeter A1) open C.B.-1 and then open the main switch DPST1.

Equalizer Bar:

Compound Generators in Parallel: Under-compounded generators also operate satisfactorily in parallel but over compounded generators will not operate satisfactorily unless their series fields are paralleled. This is achieved by connecting two negative brushes together as shown in Fig. 1.36. The conductor used to connect these brushes is generally called equalizer bar. Suppose that an attempt is made to operate the two generators in parallel without an equalizer bar. If, for any reason, the current supplied by generator 1 increases slightly, the current in its series field will increase and raise the generated voltage. This will cause generator 1 to take more load. Since total load supplied to the system is constant, the current in generator 2 must decrease and as a result its series field is weakened. Since this effect is cumulative, the generator 1 will take the entire load and drive generator 2 as a motor. After machine 2 changes from a generator to a motor, the current in the shunt field will remain in the same direction, but the current in the armature and series field will reverse. Thus the magnetizing action, of the series field opposes that of the shunt field. As the current taken by the machine 2 increases, the demagnetizing action of series field becomes greater and the resultant field becomes weaker. The resultant field will finally become zero and at that time machine 2 will be short circuited machine 1, opening the breaker of either or both machines.

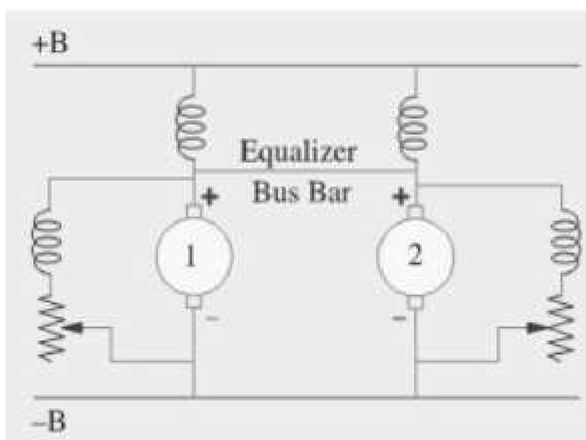


Fig. 1.36 Connection of equalizer bar in parallel connection of DC compound generator

When the equalizer bar is used, a stabilizing action exists and neither machine tends to take all the load.To consider this, suppose that current delivered by generator 1 increases. The increased current will notonly pass through the series field of generator 1 but also through the equalizer bar and series field of generator 2. Therefore, the voltage of both the machines increases and the generator 2 will take a part of the load.

Load Sharing:

The load sharing between shunt generators in parallel can be easily regulated because of their drooping characteristics. The load may be shifted from one generator to another merely by adjusting the field excitation. Let us discuss the load sharing of two generators which have unequal no-load voltages. Let E_1, E_2 = no-load voltages of the two generators R_1, R_2 = their armature resistances

V_T = common terminal voltage (Bus-bars voltage). Then

$$I_1=\frac{E_1-V_T}{R_1} \text{ and } I_2=\frac{E_2-V_T}{R_2}$$

Thus current output of the generators depends upon the values of E_1 and E_2. These values may be changed by field rheostats. The common terminal voltage (or bus-bars voltage) will depend upon (i) the emfs of individual generators and (ii) the total load current supplied. It is generally desired to keep the busbars voltage constant. This can be achieved by adjusting the field excitations of the generators operating in parallel.

Transformers

2.1 Introduction

The transformer is a device that transfers electrical energy from one electrical circuit to another electrical circuit. The two circuits may be operating at different voltage levels but always work at the same frequency. Basically transformer is an electro-magnetic energy conversion device. It is commonly used in electrical power system and distribution systems. It can change the magnitude of alternating voltage or current from one value to another. This useful property of transformer is mainly responsible for the widespread use of alternating currents rather than direct currents i.e., electric power is generated, transmitted and distributed in the form of alternating current. Transformers have no moving parts, rugged and durable in construction, thus requiring very little attention. They also have a very high efficiency as high as 99%.

2.2. Single Phase Transformer

A transformer is a static device of equipment used either for raising or lowering the voltage of an a.c. supply with a corresponding decrease or increase in current. It essentially consists of two windings, the primary and secondary, wound on a common laminated magnetic core as shown in Fig 1. The winding connected to the a.c. source is called primary winding (or primary) and the one connected to load is called secondary winding (or secondary). The alternating voltage V_1 whose magnitude is to be changed is applied to the primary.

Depending upon the number of turns of the primary (N_1) and secondary (N_2), an alternating e.m.f. E_2 is induced in the secondary. This induced e.m.f. E_2 in the secondary causes a secondary current I_2. Consequently, terminal voltage V_2 will appear across the load.

If $V_2 > V_1$, it is called a step up-transformer.

If $V_2 < V_1$, it is called a step-down transformer.

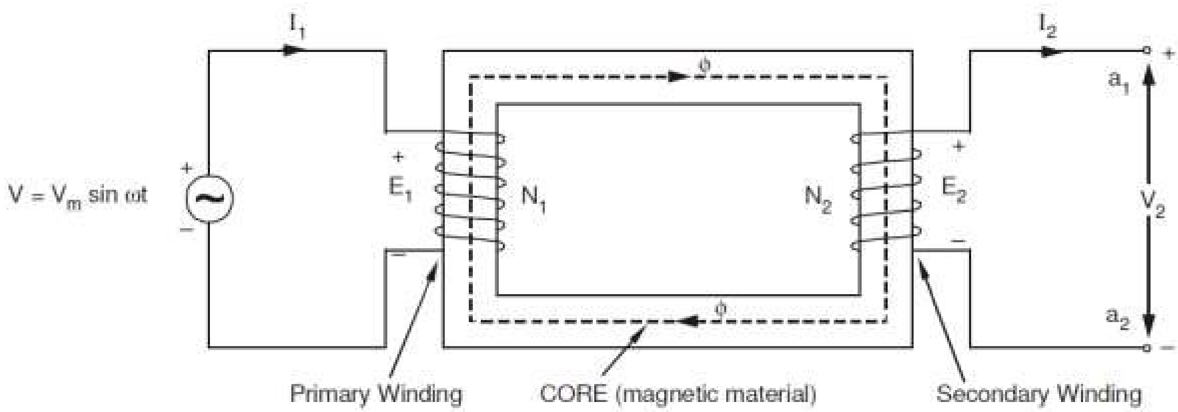


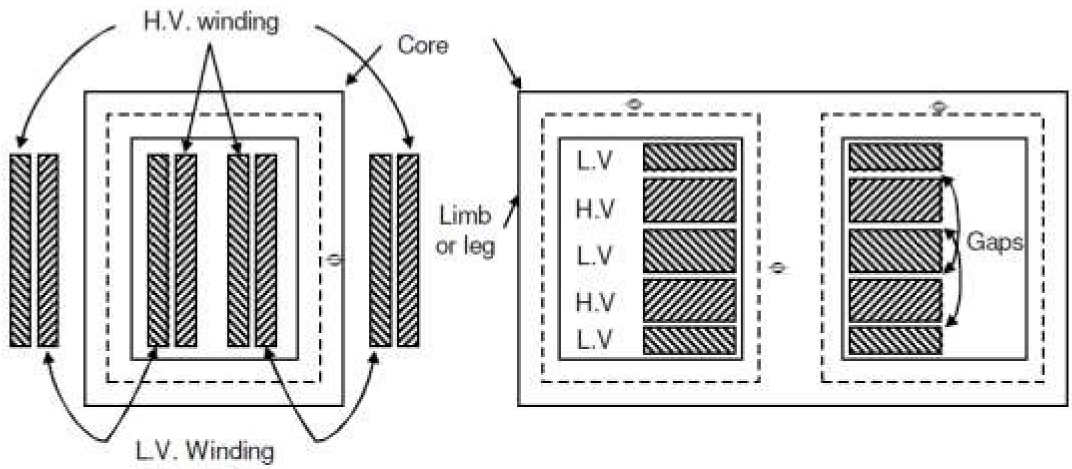
Fig. 2.1 Schematic diagram of single phase transformer

Constructional Details

Depending upon the manner in which the primary and secondary windings are placed on the core, and the shape of the core, there are two types of transformers, called (a) core type, and (b) shell type.

Core-type and Shell-type Construction

In core type transformers, the windings are placed in the form of concentric cylindrical coils placed around the vertical limbs of the core. The low-voltage (LV) as well as the high-voltage (HV) winding are made in two halves, and placed on the two limbs of core. The LV winding is placed next to the core for economy in insulation cost. Figure 2.1(a) shows the cross-section of the arrangement. In the shell type transformer, the primary and secondary windings are wound over the central limb of a three-limb core as shown in Figure 2.1(b). The HV and LV windings are split into a number of sections, and the sections are interleaved or sandwiched i.e. the sections of the HV and LV windings are placed alternately.



(a) Core Type (b) Shell Type

Fig: 2.1 Core type & shell type transformer

Core

The core is built-up of thin steel laminations insulated from each other. This helps in reducing the eddy current losses in the core, and also helps in construction of the transformer. The steel used for core is of high silicon content, sometimes heat treated to produce a high permeability and low hysteresis loss. The material commonly used for core is CRGO (Cold Rolled Grain Oriented) steel. Conductor material used for windings is mostly copper. However, for small distribution transformer aluminium is also sometimes used. The conductors, core and whole windings are insulated using various insulating materials depending upon the voltage.

Insulating Oil

In oil-immersed transformer, the iron core together with windings is immersed in insulating oil. The insulating oil provides better insulation, protects insulation from moisture and transfers the heat produced in core and windings to the atmosphere.

The transformer oil should possess the following qualities:

(a) High dielectric strength,

(b) Low viscosity and high purity,

(c) High flash point, and

(d) Free from sludge.

Transformer oil is generally a mineral oil obtained by fractional distillation of crude oil.

Tank and Conservator

The transformer tank contains core wound with windings and the insulating oil. In large transformers small expansion tank is also connected with main tank is known as conservator. Conservator provides space when insulating oil expands due to heating. The transformer tank is provided with tubes on the outside, to permits circulation of oil, which aides in cooling. Some additional devices like breather and Buchholz relay are connected with main tank. Buchholz relay is placed between main tank and conservator. It protect the transformer under extreme heating of transformer winding. Breather protects the insulating oil from moisture when the cool transformer sucks air inside. The silica gel filled breather absorbs moisture when air enters the tank. Some other necessary parts are connected with main tank like, Bushings, Cable Boxes, Temperature gauge, Oil gauge, Tappings, etc.

Principle of Operation

When an alternating voltage V1 is applied to the primary, an alternating flux ϕ is set up in the core. This alternating flux links both the windings and induces e.m.f.s E1 and E2 in them according to Faraday's laws of electromagnetic induction. The e.m.f. E1 is termed as primary e.m.f. and e.m.f. E2 is termed as secondary e.m.f.

Clearly, $$E_1 = -N_1 \frac{d\phi}{dt}$$

and $$E_2 = -N_2 \frac{d\phi}{dt}$$

$$\therefore \quad \frac{E_2}{E_1} = \frac{N_2}{N_1}$$

Note that magnitudes of E2 and E1 depend upon the number of turns on the secondary and primary respectively.

If $N_2 > N_1$, then $E_2 > E_1$ (or $V_2 > V_1$) and we get a step-up transformer. If $N_2 < N_1$, then $E_2 < E_1$ (or $V_2 < V_1$) and we get a step-down transformer.

If load is connected across the secondary winding, the secondary e.m.f. E_2 will cause a current I_2 to flow through the load. Thus, a transformer enables us to transfer a.c. power from one circuit to another with a change in voltage level.

The following points may be noted carefully:

(a) The transformer action is based on the laws of electromagnetic induction.

(b) There is no electrical connection between the primary and secondary.

(c) The a.c. power is transferred from primary to secondary through magnetic flux.

(d) There is no change in frequency i.e., output power has the same frequency as the input power.

(e) The losses that occur in a transformer are:

(a) ***core losses***—eddy current and hysteresis losses

(b) ***copper losses***—in the resistance of the windings

In practice, these losses are very small so that output power is nearly equal to the input primary power. In other words, a transformer has very high efficiency.

E.M.F. Equation of a Transformer

Consider that an alternating voltage V_1 of frequency f is applied to the primary as shown in Fig.2.3. The sinusoidal flux ϕ produced by the primary can be represented as:

$\phi = \phi_m \sin\omega t$

When the primary winding is excited by an alternating voltage V_1, it is circulating alternating current, producing an alternating flux ϕ.

ϕ - Flux

ϕ_m - maximum value of flux

N_1 - Number of primary turns

N_2 - Number of secondary turns

f - Frequency of the supply voltage

E_1 - R.M.S. value of the primary induced e.m.f

E_2 - R.M.S. value of the secondary induced e.m.f

The instantaneous e.m.f. e_1 induced in the primary is -

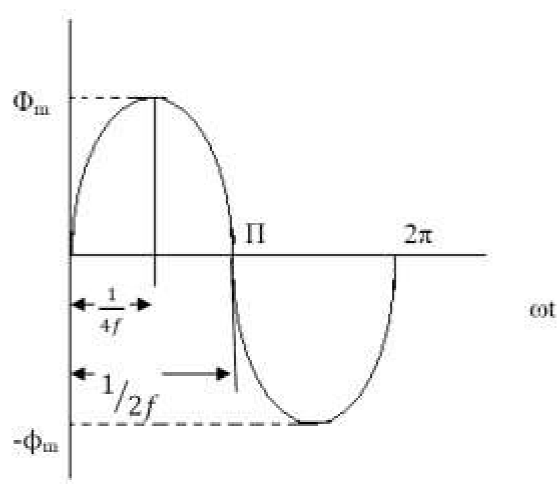


Fig. 2.3

From Faraday's law of electromagnetic induction -

$$\text{Average e.m.f per turns} = \frac{d\emptyset}{dt}$$

dϕ = change in flux
dt = time required for change in flux

The flux increases from zero value to maximum value ϕm in 1/4f of the time period that is in 1/4f seconds.

The change of flux that takes place in 1/4f seconds = $\phi_m - 0 = \phi_m$ webers

$$\frac{d\phi}{dt} = \frac{dt}{1/4f} = 4f\phi_m \ w_b/\text{sec.}$$

Since flux ϕ varies sinusoidally, the R.m.s value of the induced e.m.f is obtained by multiplying the average value with the form factor

$$\text{Form factor of a sinwave} = \frac{\text{R.m.s value}}{\text{Average value}} = 1.11$$

R.M.S Value of e.m.f induced in one turns = $4\phi_m f \times 1.11$ Volts.

$= 4.44\phi_m f$ Volts.

R.M.S Value of e.m.f induced in primary winding = $4.44\phi_m f\, N1$Volts.

R.M.S Value of e.m.f induced in secondary winding = $4.44\phi_m f\, N2$Volts.

The expression of E_1 and E_2 are called e.m.f equation of a transformer

$$\boxed{\begin{array}{l} V_1 = E_1 = 4.44\phi_m f N_1 \text{Volts.} \\ V_2 = E_2 = 4.44\phi_m f N_2 \text{Volts.} \end{array}}$$

Voltage Ratio

Voltage transformation ratio is the ratio of e.m.f induced in the secondary winding to the e.m.f induced in the primary winding.

$$\frac{E2}{E1} = \frac{4.44\phi mf\ N2}{4.44\phi mf\ N1}$$

$$\boxed{\frac{E2}{E1} = \frac{N2}{N1} = K}$$

This ratio of secondary induced e.m.f to primary induced e.m.f is known as voltage transformation ratio

$$E_2 = KE_1 \qquad \text{where } K = \frac{N2}{N1}$$

1. If $N_2 > N_1$ i.e. $K>1$ we get $E_2 > E_1$ then the transformer is called step up transformer.

2. If $N_2 < N_1$ i.e. $K<1$ we get $E_2 < E_2$ then the transformer is called step down transformer.

3. If $N_2 = N_1$ i.e. $K=1$ we get $E_2 = E_2$ then the transformer is called isolation transformer or 1:1

transformer.

Current Ratio

Current ratio is the ratio of current flow through the primary winding (I_1) to the current flowing through the secondary winding (I_2). In an ideal transformer -

Apparent input power = Apparent output power.

$$V_1I_1 = V_2I_2$$

$$\frac{I1}{I2} = \frac{V2}{V1} = \frac{N2}{N1} = K$$

Volt-Ampere Rating

i) The transformer rating is specified as the products of voltage and current (VA rating).

ii) On both sides, primary and secondary VA rating remains same. This rating is generally expressed in KVA (Kilo Volts Amperes rating).

$$\frac{V1}{V2} = \frac{I2}{I1} = K$$

$$V_1I_1 = V_2I_2$$

$$\text{KVA Rating of a transformer} = \frac{V1I1}{1000} = \frac{V2I2}{1000} \text{ (1000 is to convert KVA to VA)}$$

V_1 and V_2 are the V_t of primary and secondary by using KVA rating we can calculate I_1 and I_2 Full load current and it is safe maximum current.

$$I_1 \text{ Full load current} = \frac{KVA\ Rating\ X\ 1000}{V1}$$

$$I_1 \text{ Full load current} = \frac{KVA\ Rating\ X\ 1000}{V2}$$

Transformer on No-load

a) Ideal transformer

b) Practical transformer

a) Ideal Transformer

An ideal transformer is one that has

(i) No winding resistance

(ii) No leakage flux i.e., the same flux links both the windings

(iii) No iron losses (i.e., eddy current and hysteresis losses) in the core

Although ideal transformer cannot be physically realized, yet its study provides a very powerful tool in the analysis of a practical transformer. In fact, practical transformers have properties that approach very close to an ideal transformer.

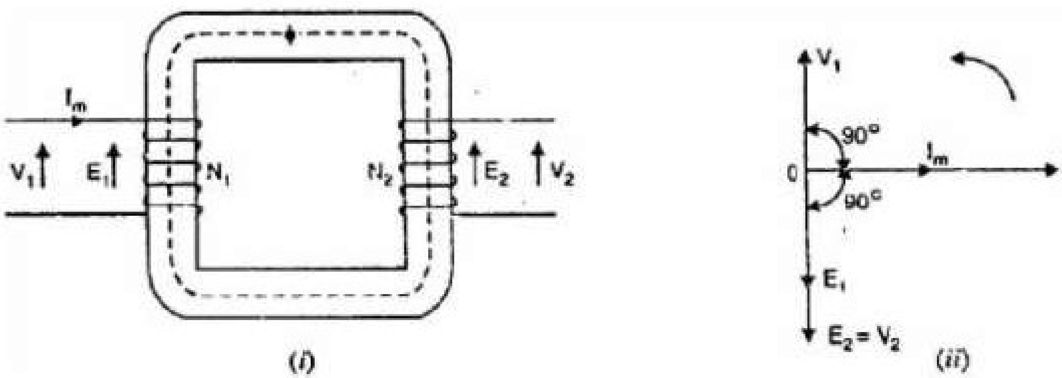


Fig: 2.4

Consider an ideal transformer on no load i.e., secondary is open-circuited as shown in *Fig.2.4 (i)*. under such conditions, the primary is simply a coil of pure inductance. When an alternating voltage V_1 is applied to the primary, it draws a small magnetizing current I_m which lags behind the applied voltage by 90°. This alternating current I_m produces an alternating flux ϕ which is proportional to and in phase with it. The alternating flux ϕ links both the windings and induces e.m.f. E_1 in the primary and e.m.f. E_2 in the secondary. The primary e.m.f. E_1 is, at every instant, equal to and in opposition to V_1 (Lenz's law). Both e.m.f.s E_1 and E_2 lag behind flux ϕ by 90°.However, their magnitudes depend upon the number of primary and secondary turns. *Fig. 2.4 (ii)* shows the phasor diagram of an ideal transformer on no load. Since flux ϕ is common to both the windings, it has been taken as the reference phasor. The primary e.m.f. E1 and secondary e.m.f. E_2 lag behind the flux ϕ by 90°. Note that E_1 and E_2 are in phase. But E_1 is equal to V_1 and 180° out of phase with it.

$$\boxed{\frac{E2}{E1} = \frac{V2}{V1} = K}$$

Phasor Diagram

i) Φ (flux) is reference

ii) Im produce ϕ and it is in phase with ϕ, V_1 Leads Im by 90˚

iii) E_1 and E_2 are in phase and both opposing supply voltage V_1, winding is purely inductive

So current has to lag voltage by 90˚.

iv) The power input to the transformer

$$P = V_1 . I_1 . \cos(90˚) \ldots\ldots\ldots\ldots (\cos 90˚ = 0)$$

$$P = 0 \text{ (ideal transformer)}$$

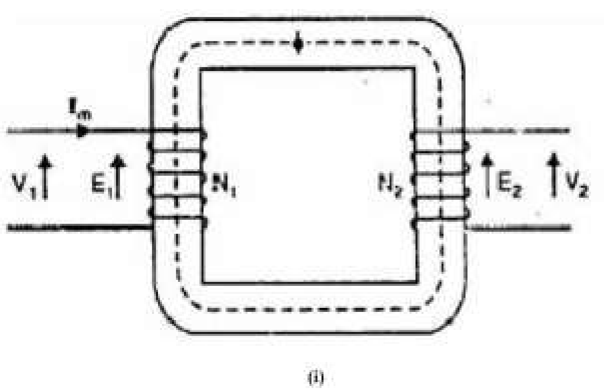
***b)i) Practical Transformer* on no load**

A practical transformer differs from the ideal transformer in many respects. The practical transformer has (i) iron losses (ii) winding resistances and (iii) Magnetic leakage

(i) Iron losses. Since the iron core is subjected to alternating flux, there occurs eddy current and hysteresis loss in it. These two losses together are known as iron losses or core losses. The iron losses depend upon the supply frequency, maximum flux density in the core, volume of the core etc. It may be noted that magnitude of iron losses is quite small in a practical transformer.

(ii) Winding resistances. Since the windings consist of copper conductors, it immediately follows that both primary and secondary will have winding resistance. The primary resistance R_1 and secondary resistance R_2 act in series with the respective windings as shown in Fig. When current flows through the windings, there will be power loss as well as a loss in voltage due to IR drop. This will affect the power factor and E_1 will be less than V_1 while V_2 will be less than E_2.

Consider a practical transformer on no load i.e., secondary on open-circuit as Shown in Fig 2.5.



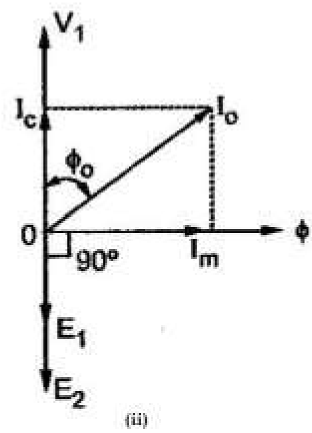


Fig: 2.5 Phasor diagram of transformer at noload

Here the primary will draw a small current I_0 to supply -

(i) the iron losses and

(ii) a very small amount of copper loss in the primary.

Hence the primary no load current I0 is not 90° behind the applied voltage V_1 but lags it by an angle ϕ_0 < 90° as shown in the phasor diagram.

No load input power, $W_0 = V_1 I_0 \cos \phi 0$

As seen from the phasor diagram in Fig.2.5 (ii), the no-load primary current I_0

(i) The component I_c in phase with the applied voltage V_1. This is known as active or working or iron loss component and supplies the iron loss and a very small primary copper loss.

$I_c = I_0 \cos \phi_0$

The component Im lagging behind V_1 by 90° and is known as magnetizing component. It is this component which produces the mutual flux ϕ in the core.

$I_m = I_0 \sin \phi_0$

Clearly, Io is phasor sum of Im and I_c,

$$I_o = \sqrt{Im^2 + Ic^2}$$

$$\text{No load P.F.,} \quad \cos\phi o = \frac{Ic}{Io}$$

The no load primary copper loss (i.e. Io^2R1) is very small and may be neglected.

Therefore, the no load primary input power is practically equal to the iron loss in the transformer i.e.,

No load input power, $W_0 = V_1 I_o \cos\phi_o = Pi$ = Iron loss

***b) ii)Practical Transformer* on Load**

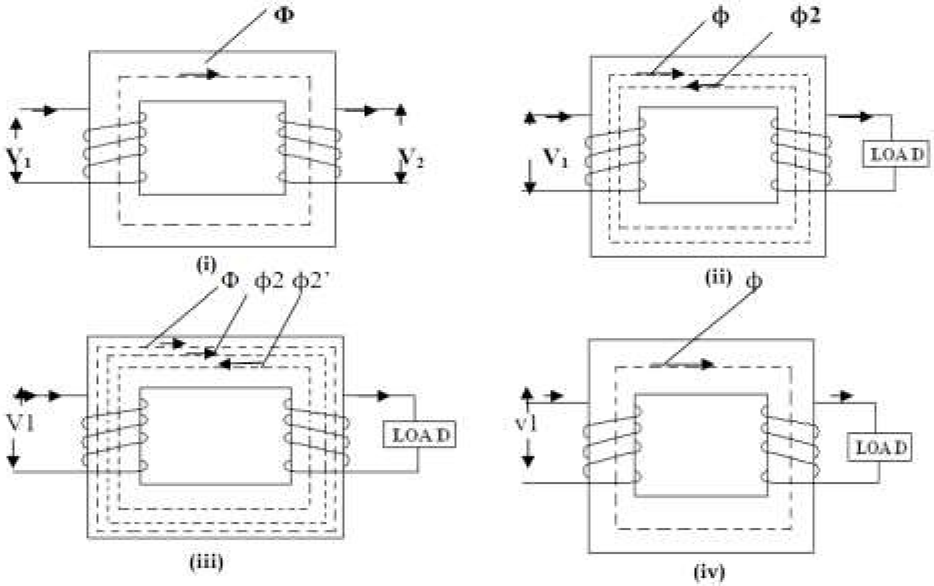


Fig: 2.6

At no load, there is no current in the secondary so that $V_2 = E_2$. On the primary side, the drops in R_1 and X_1, due to I_0 are also very small because of the smallness of I_0. Hence, we can say that at no load, $V_1 = E_1$.

i) When transformer is loaded, the secondary current I_2 is flows through the secondary winding.

ii) Already I_m magnetizing current flow in the primary winding fig. 2.6(i).

iii) The magnitude and phase of I_2 with respect to V_2 is determined by the characteristics of the load.

a) I_2 in phase with V_2 (resistive load)

b) I_2 lags with V_2 (Inductive load)

c) I_2 leads with V_2 (capacitive load)

iv) Flow of secondary current I_2 produce new Flux ϕ_2 fig.2.6 (ii)

v) Φis main flux which is produced by the primary to maintain the transformer as constant magnetising component.

vi) Φ_2 opposes the main flux ϕ, the total flux in the core reduced. It is called demagnetising Ampere-turns due to this E_1 reduced.

vii) To maintain the ϕ constant primary winding draws more current (I_2') from the supply (load component of primary) and produce ϕ_2' flux which is oppose ϕ_2 (but in same direction as ϕ), to maintain flux constant flux constant in the core fig.2.6 (iii).

viii) The load component current I_2' always neutralizes the changes in the load.

ix) Whatever the load conditions, the net flux passing through the core is approximately the same as at no-load. An important deduction is that due to the constancy of core flux at all loads, the core loss is also practically the same under all load conditions fig.2.6 (iv).

$$\boxed{\Phi_2 = \phi_2' \quad N_2I_2 = N_1I_2' \quad I_2' = \frac{N2}{N1} X I_2 = KI_2}$$

Phasor Diagram

i) Take (ϕ) flux as reference for all load

ii) The no load I_o which lags by an angle ϕo, $Io = \sqrt{Ic^2 + Im^2}$.

iii) The load component I_2', which is in anti-phase with I_2 and phase of I_2 is decided by the load.

iv) Primary current I_1 is vector sum of I_o and I_2'

$$\vec{I_1} - \vec{I_o} + \vec{I_2'}$$

$$I_1 = \sqrt{Io^2 + I2'^2}$$

a) If load is Inductive, I_2 lags E_2 by ϕ_2, shown in phasor diagram fig 2.7 (a).

b) If load is resistive, I_2 in phase with E_2 shown in phasor diagram fig. 2.7 (b).

c) If load is capacitive load, I_2 leads E_2 by ϕ_2 shown in phasor diagram fig. 2.7 (c).

For easy understanding at this stage here we assumed E_2 is equal to V_2 neglecting various drops.

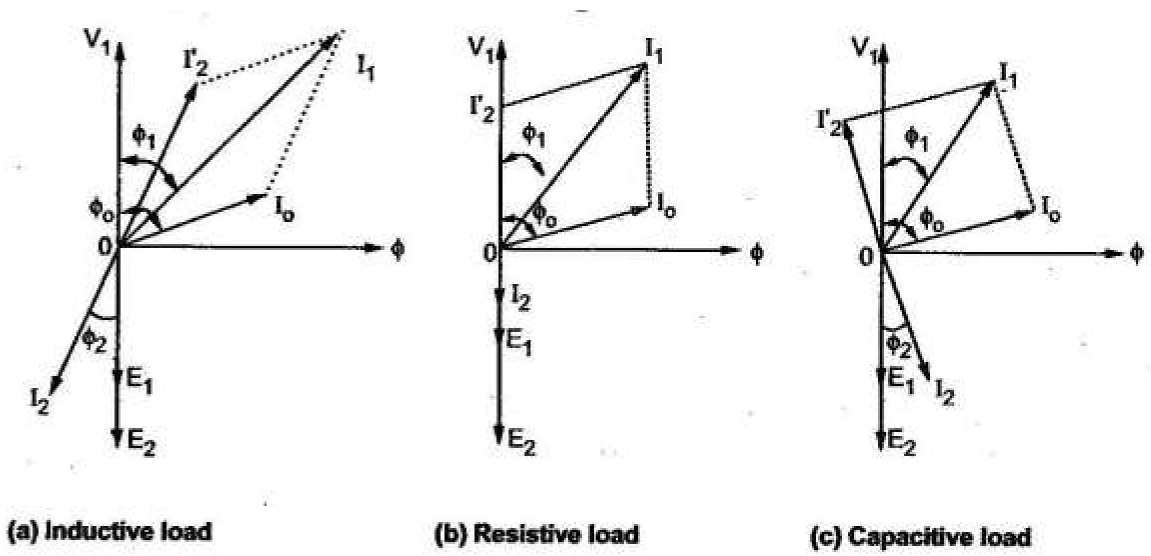


Fig: 2.7.a

$$I_1 \cong I_2'$$

Balancing the ampere – turns

$$N_1I_2' = N_1I_1 + N_2I_2$$

$$\frac{I1}{I2} = \frac{N2}{N1} = K$$

Now we going to construct complete phasor diagram of a transformer (shown in Fig: 2.7.b)

Effect of Winding Resistance

In practical transformer it process its own winding resistance causes power loss and also the voltage drop.

R_1 – primary winding resistance in ohms.

R_2 – secondary winding resistance in ohms.

The current flow in primary winding make voltage drop across it is denoted as I_1R_1 here supply voltage V_1 has to supply this drop primary induced e.m.f E_1 is the vector difference between V_1 and I_1R_1.

$$\vec{E_1} = \vec{V_1} - \overrightarrow{I_1R_1}$$

Similarly the induced e.m.f in secondary E_2, The flow of current in secondary winding makes voltage drop across it and it is denoted as I_2R_2 here E_2 has to supply this drop.

The vector difference between E_2 and I_2R_2

$\overrightarrow{V_2} = \overrightarrow{E_2} - \overrightarrow{I_2R_2}$ (Assuming as purely resistive drop here.)

Equivalent Resistance

1) It would now be shown that the resistances of the two windings can be transferred to any one of the two winding.

2) The advantage of concentrating both the resistances in one winding is that it makes calculations very simple and easy because one has then to work in one winding only.

3) Transfer to any one side either primary or secondary without affecting the performance of the transformer.

The total copper loss due to both the resistances.

$$\text{Total copper loss} = I_1^2R_1 + I_2^2R_2$$
$$= I_1^2[R_1 + \frac{I2^2}{I1^2}]$$
$$= I_1^2[R_1 + \frac{1}{K}R_2]$$

$\frac{R2}{K^2}$ is the resistance value of R_2 shifted to primary side and denoted as R_2'.

R_2' is the equivalent resistance of secondary referred to primary

$$\boxed{R_2' = \frac{R2}{K^2}}$$

Equivalent resistance of transformer referred to primary fig (ii)

$$\boxed{R_{1e} = R_1 + R_2' = R_1 + \frac{R2}{K^2}}$$

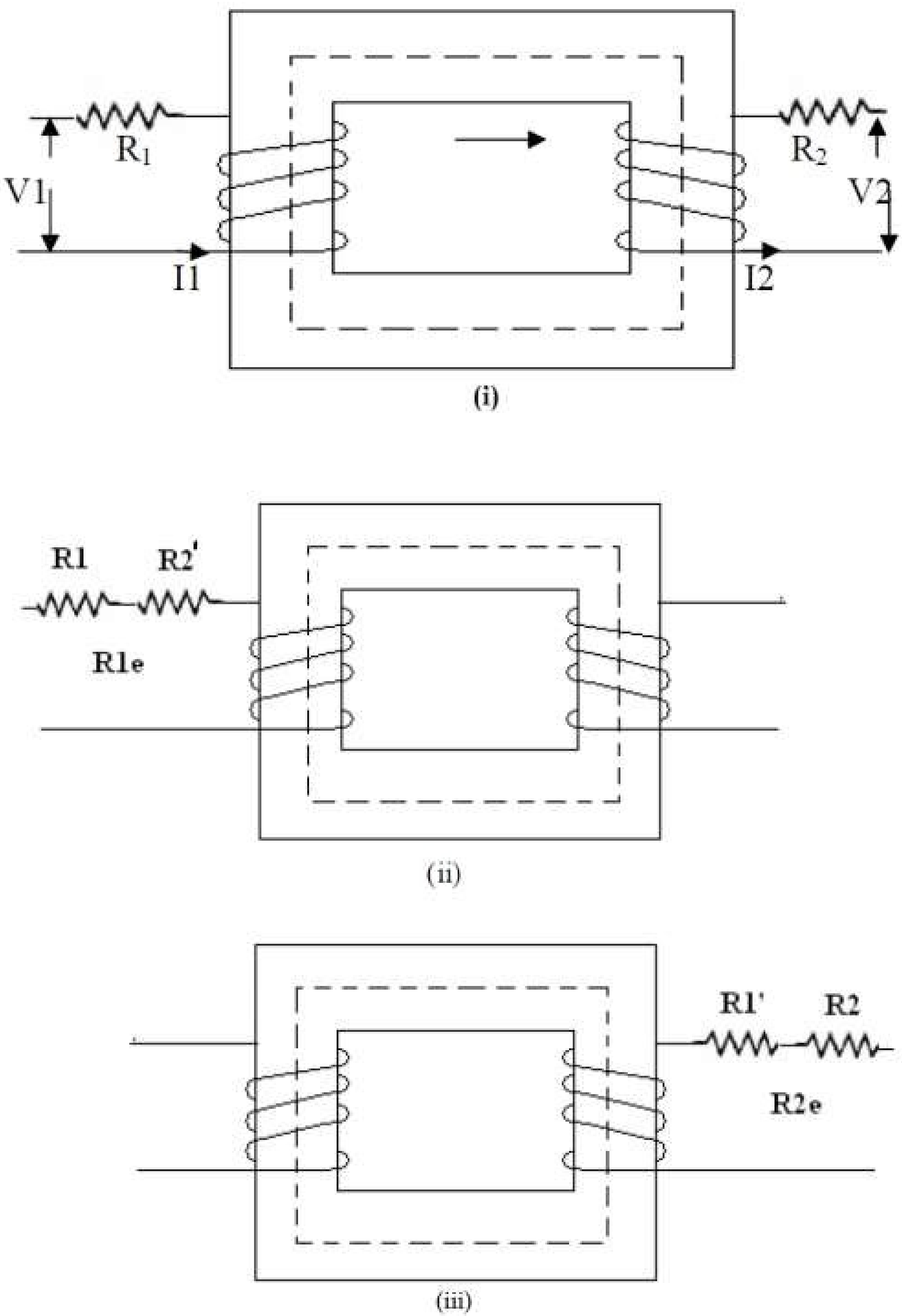


Fig:2.8

Similarly it is possible to refer the equivalent resistance to secondary winding.

Total copper loss = $I_1^2R_1 + I_2^2R_2 = I_2^2 [\frac{I1^2}{I2^2} R_1 + R_2]$

$= I_2^2[K^2R_1 + R_2]$

K^2R1 is primary resistance referred to secondary denoted as R1'.

$R_1' = K^2R_1$

Equivalent resistance of transformer referred to secondary, denoted as R2e

$$\boxed{R_{2e} = R_2 + R_1' = R_2 + K^2R_1}$$

Total copper loss = $I_2^2 R_{2e}$

Note:

Note:

i) When a resistance is to be transferred from the primary to secondary, it must be multiplied by K^2, it must be divided by K^2 while transferred from the secondary to primary.

High voltage side ➡ low current side ➡ high resistance side

Low voltage side ➡ high current side ➡ low resistance side

Effect of Leakage Reactance

i) It has been assumed that all the flux linked with primary winding also links the secondary winding. But, in practice, it is impossible to realize this condition.

ii) However, primary current would produce flux ϕ which would not link the secondary winding. Similarly, current would produce some flux ϕ that would not link the primary winding.

iii) The flux ϕ_{L1} complete its magnetic circuit by passing through air rather than around the core, as shown in fig.2.9. This flux is known as primary leakage flux and is proportional to the primary ampere – turns alone because the secondary turns do not links the magnetic circuit of ϕ_{L1}. It induces an e.m.f e_{L1} in primary but not in secondary.

iv) The flux ϕ_{L2} complete its magnetic circuit by passing through air rather than around the core, as shown in fig. This flux is known as secondary leakage flux and is proportional to the secondary ampere – turns alone because the primary turns do not links the magnetic circuit of ϕ_{L2}. It induces an e.m.f e_{L2} in secondary but not in primary.

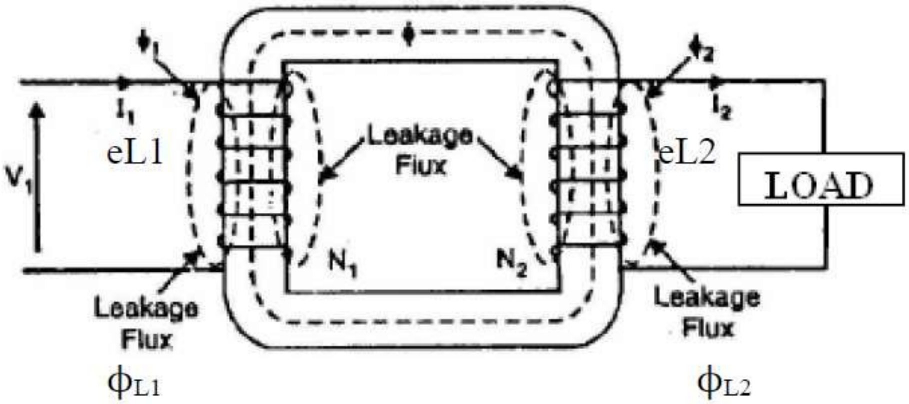


Fig: 2.9

ϕ_{L1} – primary leakage flux
ϕ_{L2} – secondary leakage flux
e_{L1} – self induced e.m.f (primary)
e_{L2} –self induced e.m.f (secondary)

Equivalent Leakage Reactance

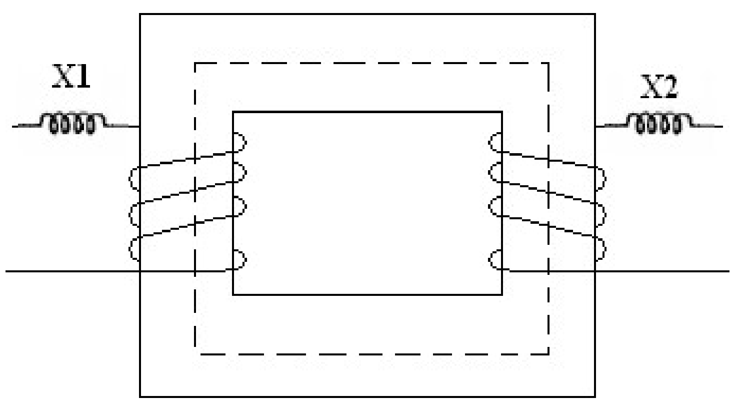


Fig: 2.10

Similarly to the resistance, the leakage reactance also can be transferred from primary to secondary. The relation through K^2 remains same for the transfer of reactance as it is studied earlier for the resistance

X_1 – leakage reactance of primary.

X_2 - leakage reactance of secondary.

Then the total leakage reactance referred to primary is X_{1e} given by

$$X_{1e} = X_1 + X_2'$$

$$X_2' = \frac{X2}{K^2}$$

The total leakage reactance referred to secondary is X_{2e} given by

$$X_{2e} = X_2 + X_1'$$
$$X_1' = K^2 X_1$$

$$\boxed{\begin{aligned} X_{1e} &= X_1 + X_2' \\ X_{2e} &= X_2 + X_1' \end{aligned}}$$

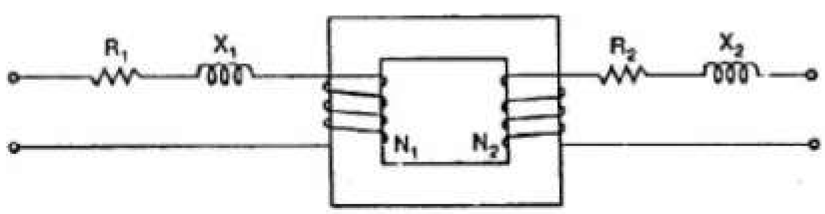
Equivalent Impedance

The transformer winding has both resistance and reactance (R_1, R_2, X_1,X_2).Thus we can say that the total impedance of primary winding isZ_1 which is,

$Z_1 = R_1 + jX_1$ ohms

On secondary winding,

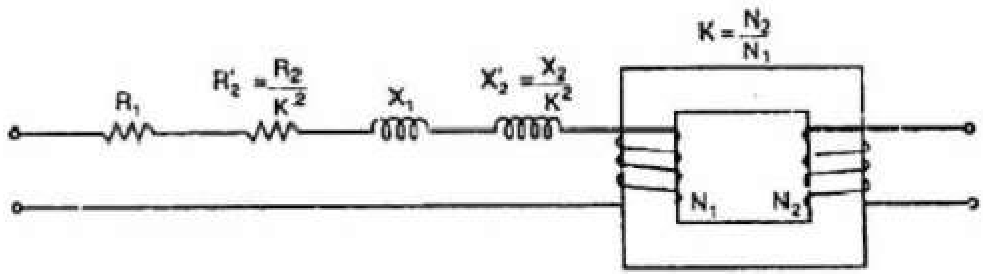
$Z_2 = R_2 + jX_2$ ohms



Individual magnitude of Z1 and Z2 are

$$Z_1 = \sqrt{R1^2 + X1^2}$$
$$Z_2 = \sqrt{R2^2 + X2^2}$$

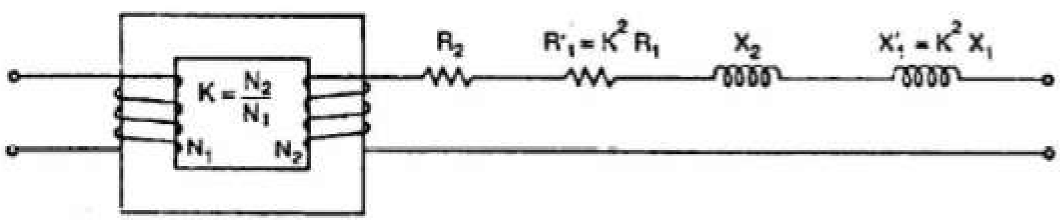
Similar to resistance and reactance, the impedance also can be referred to any one side,



Z_{1e} = total equivalent impedance referred to primary

$$Z_{1e} = R_{1e} + jX_{1e} = Z_1 + Z_2' = Z_1 + \frac{Z2}{K^2}$$

Z_{2e} = total equivalent impedance referred to secondary.



$$Z_{2e} = R_{2e} + jX_{2e} = Z_2 + Z_1' = Z_2 + K^2Z_1$$

The magnitudes of Z1e and Z2e

$$Z_1 = \sqrt{R1e^2 + X1e^2}$$
$$Z_2 = \sqrt{R2e^2 + X2e^2}$$

It can be noted that

$Z_{2e} = K^2Z_{1e}$ and $Z_{1e} = \frac{Z2e}{K^2}$

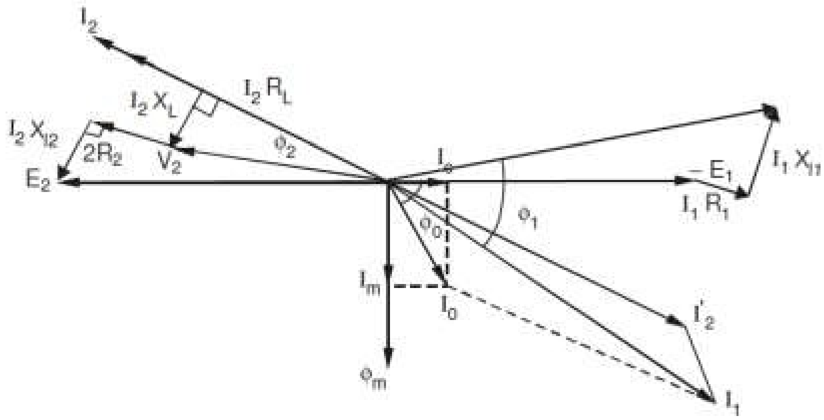
<u>Complete Phasor Diagram of a Transformer *(for Inductive Load or Lagging pf)*</u>

We now restrict ourselves to the more commonly occurring load i.e. inductive along with resistance, which has a lagging power factor.

For drawing this diagram, we must remember that

$$\overline{V_2} = \overline{E}_2 - \overline{I}_2\,(R_2 + j\,X_{l2})$$

and $$\overline{V_1} = -\,\overline{E}_1 + \overline{I}_2\,(R_1 + j\,X_{l1})$$



Equivalent Circuit of Transformer

No load equivalent circuit

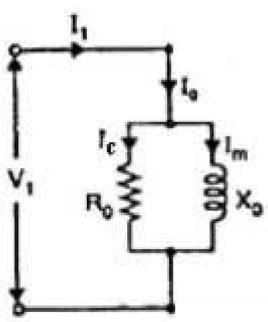


Fig:11

$I_m = I_o \sin \phi_o$ = magnetizing component
$I_c = Io \cos \phi_o$ = Active component

$R_o = \frac{V1}{Ic}$, $X_o = \frac{V1}{Im}$

i) I_m produces the flux and is assumed to flow through reactance X_o called no load reactance while I_c is active component representing core losses hence is assumed to flow through the resistance R_0

ii) Equivalent resistance is shown in fig.2.12.

iii) When the load is connected to the transformer then secondary current I_2 flows causes voltage drop across R_2 and X_2. Due to I_2, primary draws an additional current.

$$I_2' = \frac{I2}{K}$$

I_1 is the phasor addition of I_o and I_2'. This I_1 causes the voltage drop across primary resistance R_1 and reactance X_1.

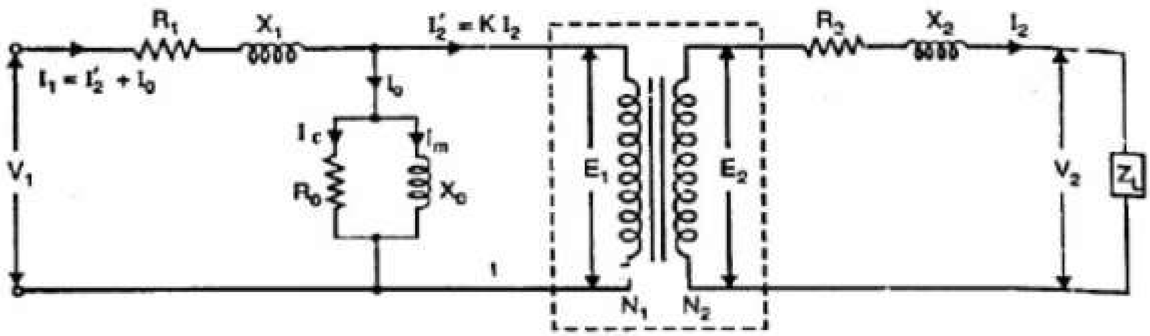


Fig: 2.12

To simplified the circuit the winding is not taken in equivalent circuit while transfer to one side.

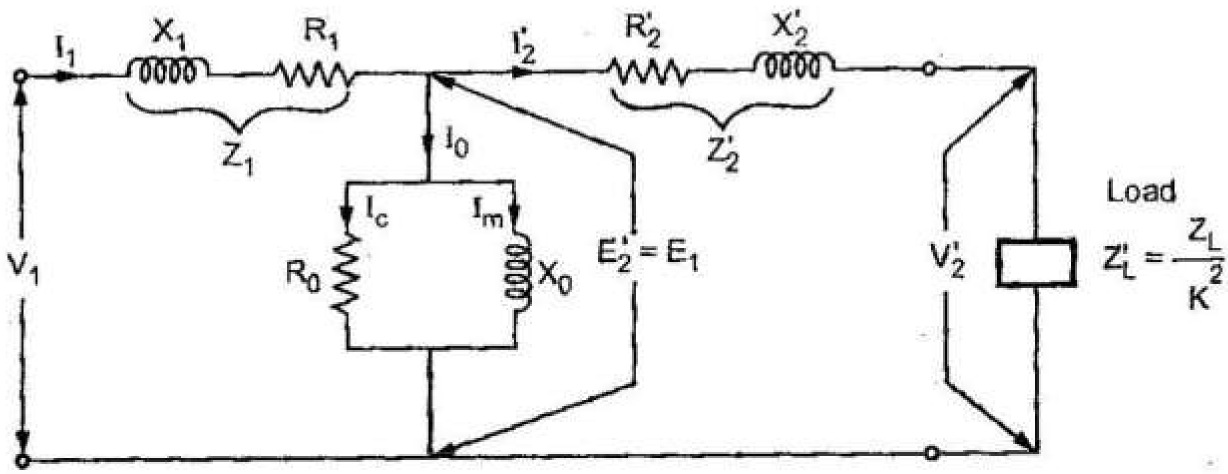


Fig: 2.13

Exact equivalent circuit referred to primary

Transferring secondary parameter to primary -

$$\boxed{R_2' = \frac{R2}{K^2},\ X_2' = \frac{X2}{K^2},\ Z_2' = \frac{Z2}{K^2},\ E_2' = \frac{E2}{K'},\ I_2' = KI_2,\ K = \frac{N2}{N1}}$$

High voltage winding ➡ low current ➡ high impedance

Low voltage winding ➡ high current ➡ low impedance

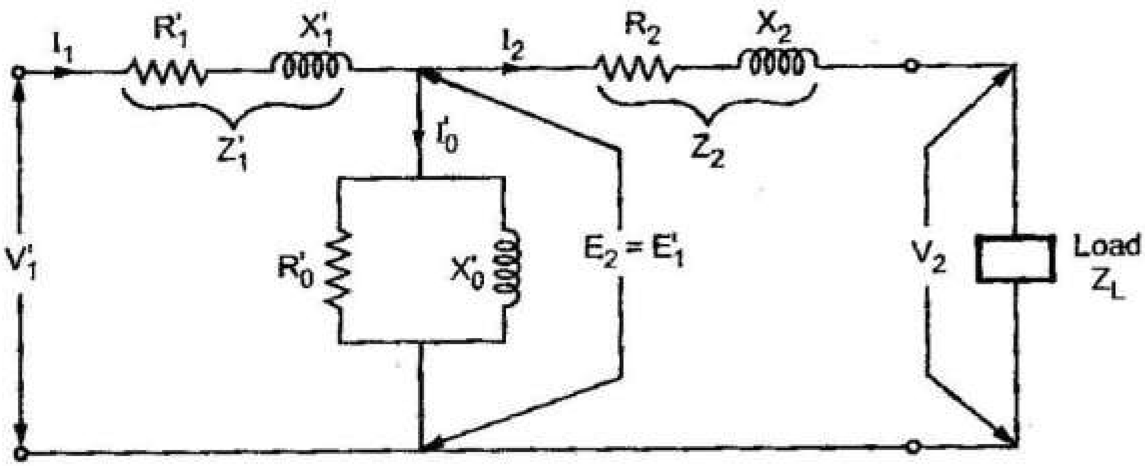


Fig: 2.14

Exact equivalent circuit referred to secondary

$$R_1' = R_1K^2,\ X_1' = K^2X_1\ ,\ E_1' = KE_1$$

$$Z_1' = K^2Z1\ ,\ I_1' = \frac{I1}{K'},\ I_o = \frac{Io}{K}$$

Now as long as no load branch i.e. exciting branch is in between Z_1 and Z_2', the impedances cannot be combined. So further simplification of the circuit can be done. Such circuit is called approximate equivalent circuit.

Approximate Equivalent Circuit

i) To get approximate equivalent circuit, shift the no load branch containing R_o and X_o to the left of R_1 and X_1.

ii) By doing this we are creating an error that the drop across R_1 and X_1 to I_o is neglected due to this circuit because simpler.

iii) This equivalent circuit is called approximate equivalent circuit Fig: 2.15 & Fig: 2.16.

In this circuit new R_1 and R_2' can be combined to get equivalent circuit referred to primary R_{1e},similarly

X_1 and X_2' can be combined to get X_{1e}.

$$R_{1e} = R_1 + R_2' = R_1 + \frac{R2}{K^2}$$

$$X_{1e} = X_1 + X_2' = X_1 + \frac{X2}{K^2}$$

$$Z_{1e} = R_{1e} + jX_1E, \qquad R_o = \frac{V1}{Ic}, \quad \text{and } X_o = \frac{V1}{Im}$$

$$I_c = I_o \cos\phi_o, \quad \text{and } I_m = I_o \sin\phi_o$$

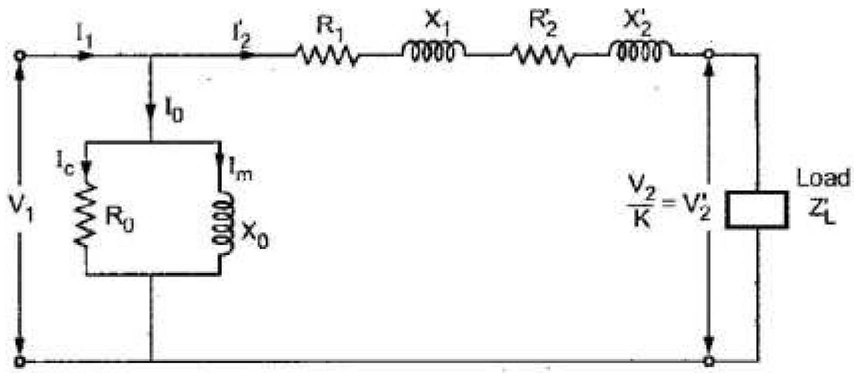


Fig:2.15Approximate equivalent circuit referred to primary

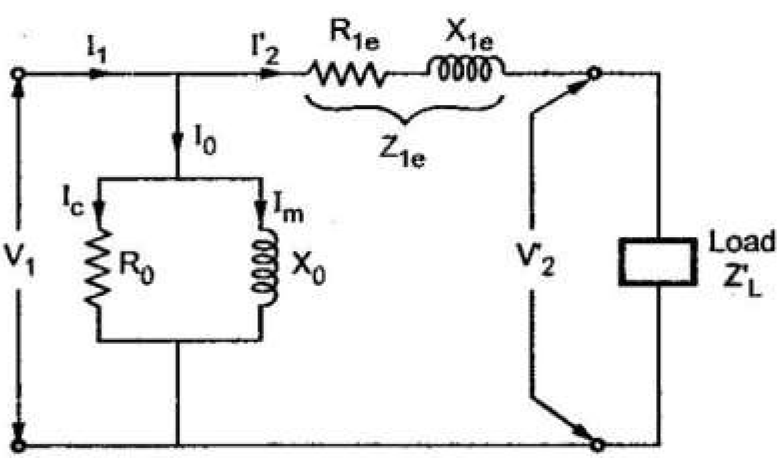
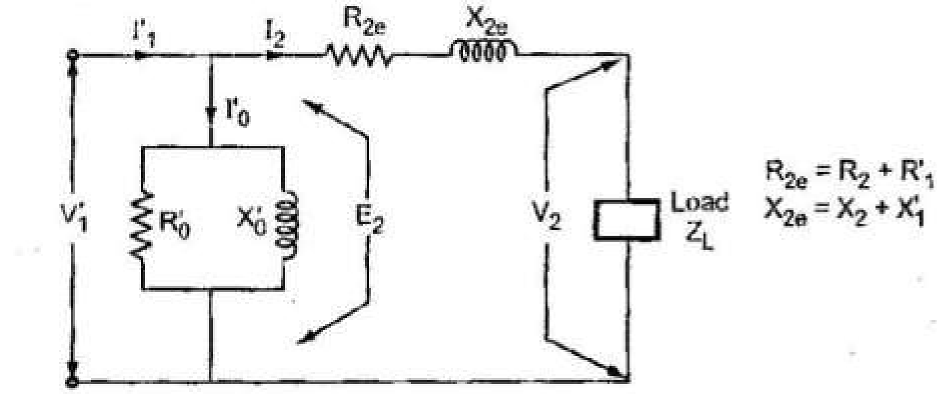


Fig:2.16Simplified equivalent circuit

Approximate Voltage Drop in a Transformer



$$\vec{E_2} = \overrightarrow{I_2R_{2e}} + \overrightarrow{I_2X_{2e}} + \vec{V_2} = \vec{V_2} + \vec{I_2}(R_{2e} + jX_{2e}), \quad \vec{E_2} = \vec{V_2} + \overrightarrow{I_2Z_{2e}}$$

Fig. 2.17

Primary parameter is referred to secondary there are no voltage drop in primary. When there is no load, $I_2 = 0$ and we get no load terminal voltage drop in

$V_{2o} = E_2$ = no load terminal voltage

V_2 = terminal voltage on load

For Lagging P.F.

i) The current I_2 lags V_2 by angle ϕ_2

ii) Take V_2 as reference

iii) I_2R_{2e} is in phase with I_2 while $I_2 X_{2e}$ leads I_2 by 90˚

iv) Draw the circle with O as centre and OC as radius cutting extended OA at M. as OA = V_2 and now OM = E_2.

v) The total voltage drop is AM = I_2Z_{2e}.

vi) The angle α is practically very small and in practice M&N are very close to each other. Due to this the approximate voltage drop is equal to AN instead of AM

AN – approximate voltage drop

To find AN by adding AD& DN

$AD = AB \cos\phi = I_2R_{2e} \cos\phi$

$DN = BL \sin\phi = I_2X_{2e} \sin\phi$

$AN = AD + DN = I_2R_{2e} \cos\phi_2 + I_2X_{2e} \sin\phi_2$

Assuming: $\phi_2 = \phi_1 = \phi$

Approximate voltage drop = $I_2R_{2e} \cos\phi + I_2X_{2e} \sin\phi$ (referred to secondary)

Similarly: Approximate voltage drop = $I_1R_{1e} \cos\phi + I_1X_{1e} \sin\phi$ (referred to primary)

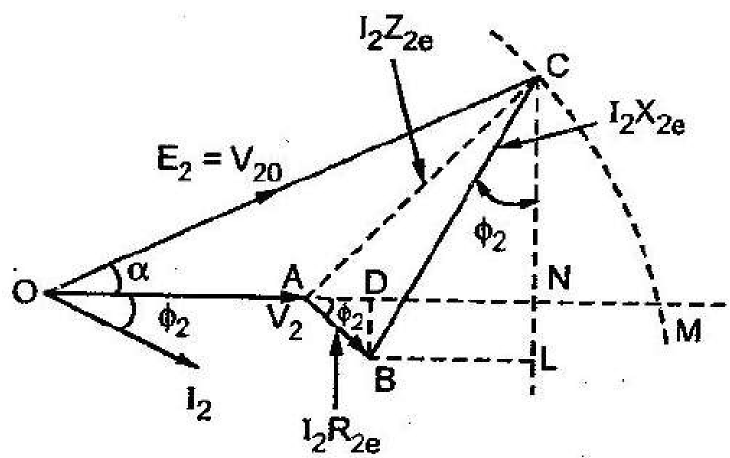


Fig:2.18

For Leading P.F Loading

I_2 leads V_2 by angle ϕ_2

Approximate voltage drop = $I_2R_{2e} \cos\phi - I_2X_{2e} \sin\phi$ (referred to secondary)

Similarly: Approximate voltage drop = $I_1R_{1e} \cos\phi - I_1X_{1e} \sin\phi$ (referred to primary)

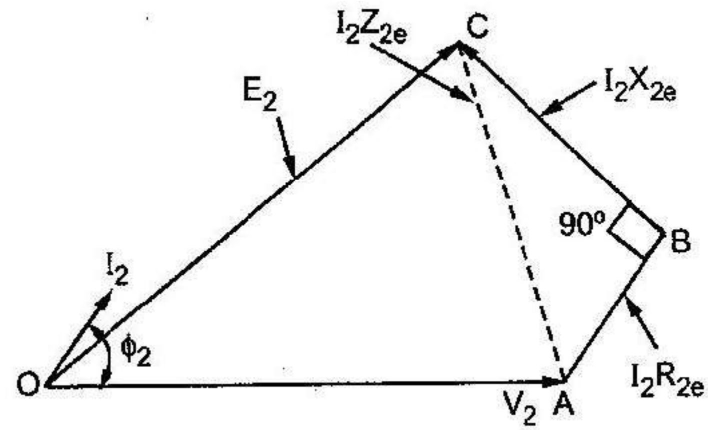
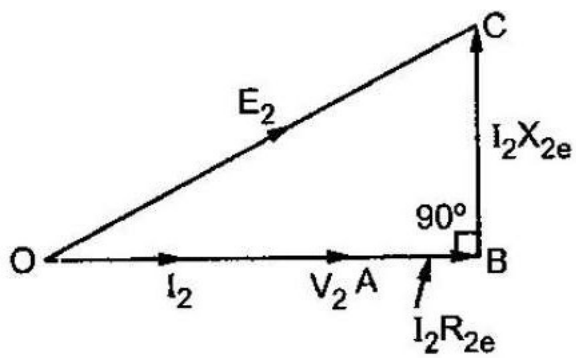


Fig: 2.19

For Unity P.F. Loading

Approximate voltage drop = I_2R_{2e} (referred to secondary)

Similarly: Approximate voltage drop = I_1R_{1e} (referred to primary)



Cosϕ = 1
Sinϕ = 0

Fig: 2.20

Approximate voltage drop = $E_2 - V_2$

$= I_2R_{2e}\cos\phi \pm I_2X_{2e}\sin\phi$ (referred to secondary)

$= I_1R_{1e}\cos\phi \pm I_1X_{1e}\sin\phi$ (referred to primary)

Losses in a Transformer

The power losses in a transformer are of two types, namely;

1. Core or Iron losses

2. Copper losses

These losses appear in the form of heat and produce (i) an increase in Temperature and (ii) a drop in efficiency.

Core or Iron losses (Pi)

These consist of hysteresis and eddy current losses and occur in the transformer core due to the alternating flux. These can be determined by open-circuit test.

Hysteresis loss = k_h f $Bm^{1.6}$ watts /m^3

K_h – hysteresis constant depend on material

f - Frequency

B_m – maximum flux density

Eddy current loss = K_e f^2 $B_m^2 t^2$ watts /m^3

Ke – eddy current constant

t - Thickness of the core

Both hysteresis and eddy current losses depend upon

(i) Maximum flux density B_m in the core

(ii) Supply frequency f. Since transformers are connected to constant-frequency, constant voltage supply, both f and Bm are constant. Hence, core or iron losses are practically the same at all loads.

Iron or Core losses, Pi = Hysteresis loss + Eddy current loss = Constant losses (Pi)

The hysteresis loss can be minimized by using steel of high silicon content .Whereas eddy current loss can be reduced by using core of thin laminations.

Copper losses (P_{cu})

These losses occur in both the primary and secondary windings due to their ohmic resistance. These can be determined by short-circuit test. The copper loss depends on the magnitude of the current flowing through the windings.

$$\text{Total copper loss} = I_1^2R_1 + I_2^2R_2 = I_1^2(R_1 + R_2') = I_2^2\,(R_2 + R_1')$$

$$\boxed{\text{Total loss} = \text{iron loss} + \text{copper loss} = Pi + Pcu}$$

Efficiency of a Transformer

Like any other electrical machine, the efficiency of a transformer is defined as the ratio of output power (in watts or kW) to input power (watts or kW) i.e.

Power output = power input – Total losses

Power input = power output + Total losses

= power output + Pi + P_{cu}

$$\text{Efficiency} = \frac{power\ output}{power\ input}$$

$$\text{Efficiency} = \frac{power\ output}{power\ input + Pi + Pcu}$$

Power output = $V_2I_2 \cos\phi$, $\cos\phi$ = load power factor

Transformer supplies full load of current I_2 and with terminal voltage V_2

Pcu = copper losses on full load = $I_2^2\,R_{2e}$

$$\text{Efficiency} = \frac{\text{V2I2 cos}\phi}{\text{V2I2 cos}\phi + Pi + I2^2R2e}$$

V_2I_2 = VA rating of a transformer

$$\text{Efficiency} = \frac{\text{(VA rating)X cos}\phi}{\text{(VA rating)X cos}\phi + Pi + I2^2R2e}$$

$$\boxed{\%\ \text{Efficiency} = \frac{\text{(VA rating)X cos}\phi}{\text{(VA rating)X cos}\phi + Pi + I2^2R2e}\ \text{X 100}}$$

This is full load efficiency and I_2 = full load current.

We can now find the full-load efficiency of the transformer at any p.f. without actually loading the transformer.

$$\text{Full load Efficiency} = \frac{(\text{Full load VA rating})\text{X}\cos\phi}{(\text{Full load VA rating})\text{X}\cos\phi + Pi + I2^2R2e}$$

Also for any load equal to n x full-load,

Corresponding total losses = $P_i + n^2P_{Cu}$

$$n = \text{fractional by which load is less than full load} = \frac{actual\ load}{full\ load}$$

$$n = \frac{half\ load}{fullload} = \frac{(\frac{1}{2})}{1} = 0.5$$

$$\text{Corresponding (n) \% Efficiency} = \frac{n(\text{VA rating})\text{X}\cos\phi}{n(\text{VA rating})\text{X}\cos\phi + Pi + n^2Pcu} \text{X } 100$$

Condition for Maximum Efficiency

Voltage and frequency supply to the transformer is constant the efficiency varies with the load. As load increases, the efficiency increases. At a certain load current, it loaded further the efficiency start decreases as shown in fig. 2.21.

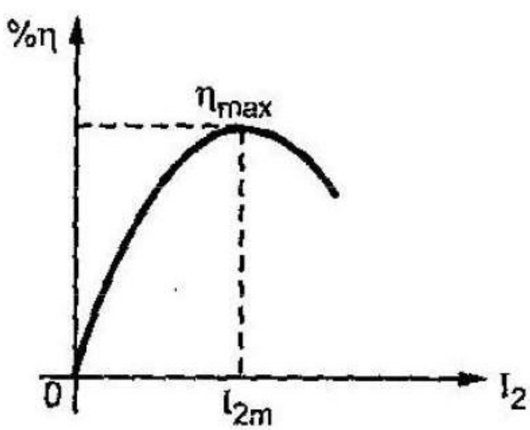


Fig: 2.21

The load current at which the efficiency attains maximum value is denoted as I_{2m} a n d maximum efficiency is denoted as ηmax, now we find -

(a) condition for maximum efficiency

(b) load current at which η_{max} occurs

(c) KVA supplied at maximum efficiency

Considering primary side,

Load output = $V_1 I_1 \cos\phi 1$

Copper loss = $I_1^2 R_{1e}$ or $I2^2 R2e$

Iron loss = hysteresis + eddy current loss = Pi

$$\text{Efficiency} = \frac{V1I1\cos\phi1 - \text{losses}}{V1I1\cos\phi1} = \frac{V1I1\cos\phi1 - I1^2R1e + Pi}{V1I1\cos\phi1}$$

$$= 1 - \frac{I1R1e}{V1I21\cos\phi1} = \frac{Pi}{V1I1\cos\phi1}$$

Differentiating both sides with respect to I2, we get

$$\frac{d\eta}{dI2} = 0 - \frac{R1e}{V1\cos\phi1} = \frac{Pi}{V1I1^2\cos\phi1}$$

For η to be maximum, $\frac{d\eta}{dI2} = 0$. Hence, the above equation becomes

$$\frac{R1e}{V1\cos\phi1} = \frac{Pi}{V1I1^2\cos\phi1} \text{ or } Pi = I_1^2R_{1e}$$

P_{cu} loss = P_i iron loss

The output current which will make P_{cu} loss equal to the iron loss. By proper design, it is possible to make the maximum efficiency occur at any desired load.

Load current I2m at maximum efficiency

For η_{max} $I_2^2 R_{2e} = P_i$ but $I_2 = I_{2m}$

$I^2_{2m} R_{2e} = Pi$ $\qquad\qquad I_{2m} = \sqrt{\frac{Pi}{R2e}}$

This is the load current at η_{max}

$(I_2)F.L$ = full load current

$$\frac{I2m}{(I2)F.L} = \frac{1}{(I2)FL}\sqrt{\frac{Pi}{R2e}}$$

$$\frac{I2m}{(I2)F.L} = \sqrt{\frac{Pi}{[(I2)F.L]^2 R2e}} = \sqrt{\frac{Pi}{[Pcu]F.L}}$$

$$\boxed{I_{2m} = (I_2)\ F.L.\sqrt{\frac{Pi}{[Pcu]F.L}}}$$

This is the load current at η_{max} in terms of full load current

KVA Supplied at Maximum Efficiency

For constant V_2 the KVA supplied is the function of load current.

$$\text{KVA at } \eta_{max} = I_{2m} V_2 = V_2(I_2)\text{F.L. X } \sqrt{\frac{Pi}{[Pcu]\text{F.L}}}$$

$$\boxed{\text{KVA at } \eta_{max} = (\text{KVA rating}) \text{ X } \sqrt{\frac{Pi}{[Pcu]\text{F.L}}}}$$

Substituting condition for η_{max} in the expression of efficiency, we can write expression for η_{max} as ,

as Pcu = Pi

$$\% \ \eta_{max} = \frac{V2 \ I2m \cos\phi}{V2 \ I2m\cos\phi + 2Pi} \text{ X } 100$$

$$\% \ \eta_{max} = \frac{KVA \ for \eta \max \cos\varphi}{KVA \ for \eta \max \cos\varphi + 2Pi}$$

12.8.4 All Day Efficiency (Energy Efficiency)

In electrical power system, we are interested to find out the all-day efficiency of any transformer because the load at transformer is varying in the different time duration of the day. So all day efficiency is defined as the ratio of total energy output of transformer to the total energy input in 24 hours.

$$\text{All day efficiency} = \frac{\text{kWh output during a day}}{\text{kWh input during the day}}$$

Here, kWh is kilowatt hour.

12.9 Testing of Transformer

The testing of transformer means to determine efficiency and regulation of a transformer at any load and at any power factor condition.

There are two methods

i) Direct loading test

ii) Indirect loading test

a. Open circuit test

b. Short circuit test

i) Load test on transformer

This method is also called as direct loading test on transformer because the load is directly connected to the transformer. We required various meters to measure the input and output reading while change the load from zero to full load. Fig. 2.22 shows the connection of transformer for direct load test. The primary is connected through the variac to change the input voltage as we required. Connect the meters as shown in the figure below.

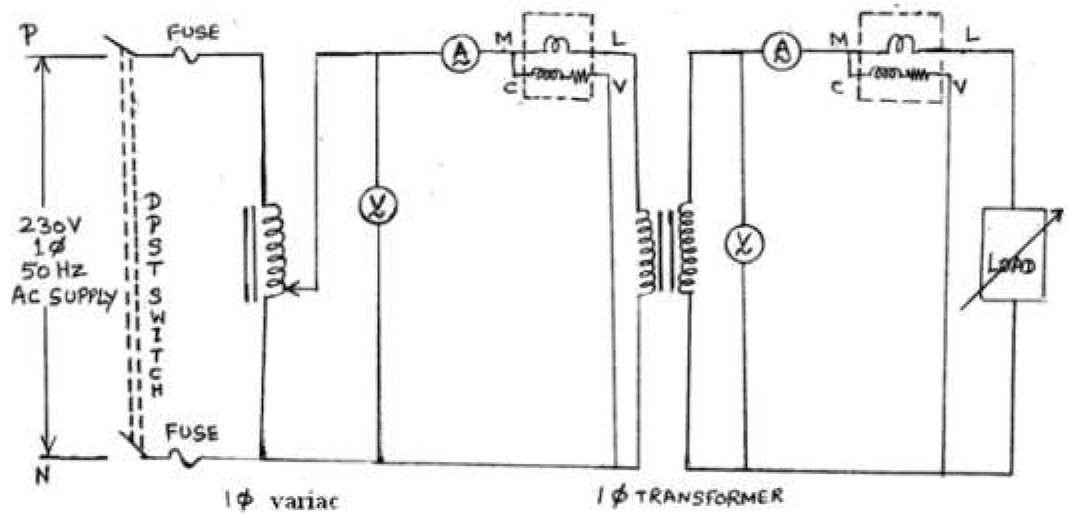


Fig: 2.22

The load is varied from no load to full load in desired steps. All the time, keep primary voltage V_1 constant at its rated value with help of variac and tabulated the reading. The first reading is to be noted on no load for which $I_2 = 0$ A and $W_2 = 0$W.

Calculation

From the observed reading

W_1 = input power to the transformer

W_2 = output power delivered to the load

$$\% \eta = \frac{W2}{W1} \text{ X } 100$$

The first reading is no load so $V2 = E_2$

The regulation can be obtained as

$$\% \text{ R} = \frac{E2-V2}{V2} \text{ X } 100$$

The graph of % η and % R on each load against load current I_L is plotted as shown in fig. 2.23.

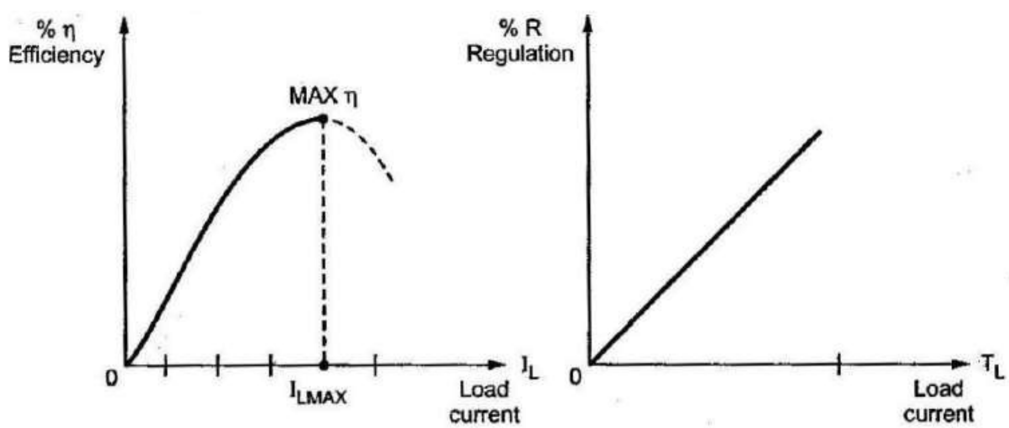


Fig: 2.23

Advantages:

1) This test enables us to determine the efficiency of the transformer accurately at any load.

2) The results are accurate as load is directly used.

Disadvantages:

1) There are large power losses during the test.

2) Load not avail in lab while test conduct for large transformer.

ii) a. Open-Circuit or No-Load Test

This test is conducted to determine the iron losses (or core losses) and parameters R_0 and X_0 of the transformer. In this test, the rated voltage is applied to the primary (usually low-voltage winding) while the secondary is left open circuited. The applied primary voltage V_1 is measured by the voltmeter, the no load current I_0 by ammeter and no-load input power W_0 by wattmeter as shown in Fig.2.24.a. As the normal rated voltage is applied to the primary, therefore, normal iron losses will occur in the transformer core. Hence wattmeter will record the iron losses and small copper loss in the primary. Since no-load current I_0 is very small (usually 2-10 % of rated current). Cu losses in the primary under no-load condition are negligible as compared with iron losses. Hence, wattmeter reading practically gives the iron losses in the transformer. It is reminded that iron losses are the same at all loads.

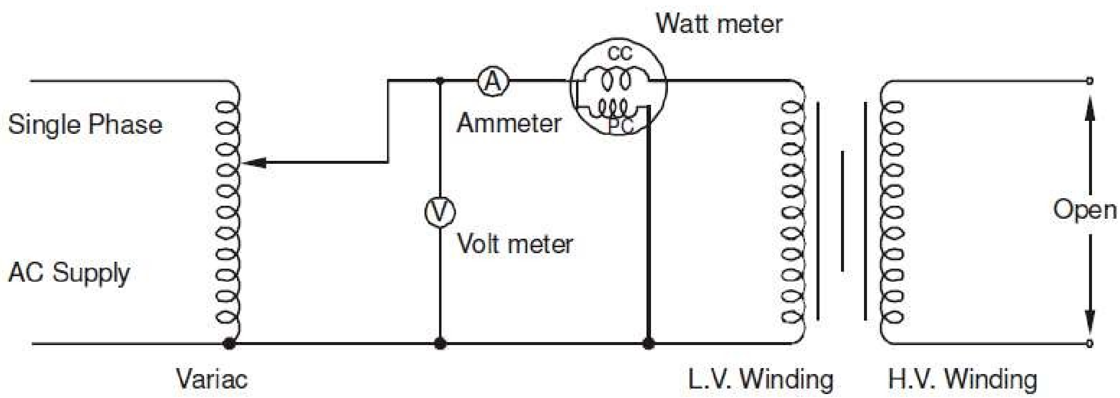


Fig: 2.24.a

Iron losses, P_i = Wattmeter reading = W_0

No load current = Ammeter reading = I_0

Applied voltage = Voltmeter reading = V_1

Input power, $W_0 = V_1 I_0 \cos\phi_0$

No - load p.f., $\cos\phi = \frac{Wo}{Vo\,Io}$ = no load power factor

$I_m = I_o \sin\phi_o$ = magnetizing component

$I_c = I_o \cos\phi_o$ = Active component

$$R_o = \frac{Vo}{Ic}\ \Omega, \quad X_o = \frac{Vo}{Im}\ \Omega$$

Under no load conditions the PF is very low (near to 0) in lagging region. By using the above data we can draw the equivalent parameter shown in Figure 2.24.b.

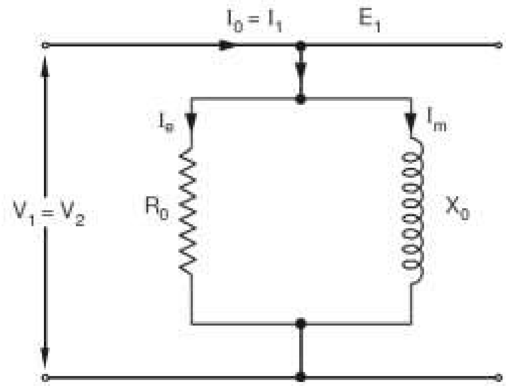


Fig: 2.24.b

Thus open-circuit test enables us to determine iron losses and parameters R_0 and X_0 of the transformer.

ii) b. Short-Circuit or Impedance Test

This test is conducted to determine R_{1e} (or R_{2e}), X_{1e} (or X_{2e}) and full-load copper losses of the transformer. In this test, the secondary (usually low-voltage winding) is short-circuited by a thick conductor and variable low voltage is applied to the primary as shown in Fig.2.25. The low input voltage is gradually raised till at voltage V_{SC}, full-load current I_1 flows in the primary. Then I_2 in the secondary also has full-load value since $I_1/I_2 = N_2/N_1$. Under such conditions, the copper loss in the windings is the same as that on full load. There is no output from the transformer under short-circuit conditions. Therefore, input power is all loss and this loss is almost entirely copper loss. It is because iron loss in the core is negligibly small since the voltage V_{SC} is very small. Hence, the wattmeter will practically register the full load copper losses in the transformer windings.

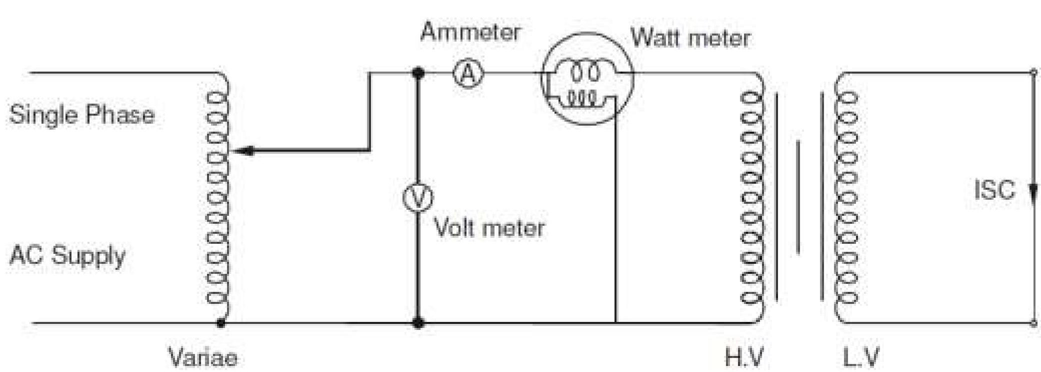


Fig: 2.25.a

Full load Cu loss, PC = Wattmeter reading = W_{sc}
Applied voltage = Voltmeter reading = V_{SC}
F.L. primary current = Ammeter reading = I_1

$Pcu = I_1^2R_1 + I_1^2 R_2' = I_1^2 R_{1e}$, $R_{1e} = \dfrac{Pcu}{I1^2}$

Where R_{1e} is the total resistance of transformer referred to primary.

Total impedance referred to primary, $Z1e = \sqrt{Z1e^2 - R1e^2}$,

short – circuit P.F, $\cos\Phi = \dfrac{Pcu}{Vsc\, I1}$ Thus short-circuit lest gives full-load Cu loss, R1e and X1e.

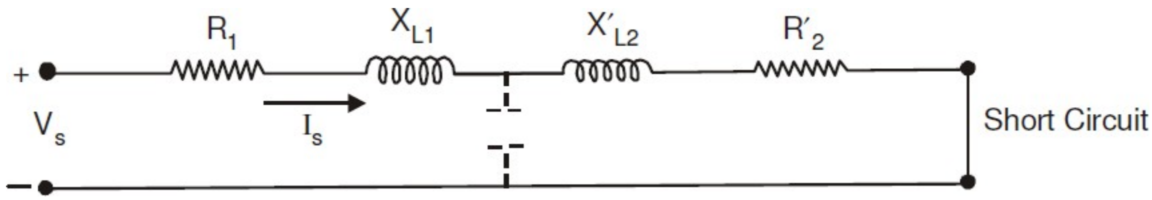


Fig: 2.25.b

From fig: 2.25.b we can calculate,

equivalent resistance $R_{eq} = \dfrac{W_s}{I_s^2} = R_1 + R_2'$

and equivalent impedance $Z_{eq} = \dfrac{V_s}{I_s}$

So we calculate equivalent reactance

$$X_{eq} = \sqrt{Z_{eq}^2 - R_{eq}^2} = X_{L1} + X'_{L2}$$

These R_{eq} and X_{eq} are equivalent resistance and reactance of both windings referred in HV side. These are know as equivalent circuit resistance and reactance.

Voltage Regulation of Transformer

Under no load conditions, the voltage at the secondary terminals is E_2 and

$$E_2 \approx V_1 \cdot \frac{N_2}{N_1}$$

(This approximation neglects the drop R_1 and X_{l1} due to small no load current). As load is applied to the transformer, the load current or the secondary current increases. Correspondingly, the primary current

I_1 also increases. Due to these currents, there is a voltage drop in the primary and secondary leakage reactances, and as a consequence the voltage across the output terminals or the load terminals changes. In quantitative terms this change in terminal voltage is called Voltage Regulation.

Voltage regulation of a transformer is defined as the drop in the magnitude of load voltage (or secondary terminal voltage) when load current changes from zero to full load value. This is expressed as a fraction of secondary rated voltage.

$$\text{Regulation} = \frac{\text{Secondary terminal voltage at no load} - \text{Secondary terminal voltage at any load}}{\text{Secondary rated voltage}}$$

The secondary rated voltage of a transformer is equal to the secondary terminal voltage at no load (i.e. E_2), this is as per IS.

Voltage regulation is generally expressed as a percentage.

$$\text{Percent voltage regulation (\% VR)} = \frac{E_2 - V_2}{E_2} \times 100.$$

Note that E_2, V_2 are magnitudes, and not phasor or complex quantities. Also note that voltage regulation depends not only on load current, but also on its power factor. Using approximate equivalent circuit referred to primary or secondary, we can obtain the voltage regulation. From approximate equivalent circuit referred to the secondary side and phasor diagram for the circuit.

$$E_2 = V_2 + I_2\ r_{eq}\ \cos\phi_2 \pm I_2\ x_{eq}\ \sin\phi_2$$

where $r_{eq} = r_2 + r_1^1$ (referred to secondary) $x_e = x_2 + x_1^1$ (+ sign applies lagging power factor load and – sign applies to leading pf load).

So $$\frac{E_2 - V_2}{E_2} = \frac{I_2\ r_{eq}\ \cos\phi_2 \pm I_2\ x_{eq}\ \sin\phi_2}{E_2}$$

$$\frac{E_2 - V_2}{E_2} = \frac{I_2\ r_{eq}}{E_2}\cos\phi_2 \pm \frac{I_2\ x_{eq}}{E_2}\sin\phi_2$$

% Voltage regulation = (% resistive drop) $\cos\phi_2 \pm$ (% reactive drop) $\sin\phi_2$.

Ideally voltage regulation should be zero.

Auto-transformers

The transformers we have considered so far are two-winding transformers in which the electrical circuit connected to the primary is electrically isolated from that connected to the secondary. An auto-transformer does not provide such isolation, but has economy of cost combined with increased efficiency. Fig.2.26 illustrates the auto-transformer which consists of a coil of N_A turns between terminals 1 and 2, with a third terminal 3 provided after N_B turns. If we neglect coil resistances and leakage fluxes, the flux linkages of the coil between 1 and 2 equals $N_A\ \phi m$ while the portion of coil between 3 and 2 has a flux linkage $N_B\ \phi m$. If the induced voltages are designated as E_A and E_B, just as in a two winding transformer,

$$\frac{\overline{E}_A}{\overline{E}_B} = \frac{N_A}{N_B}$$

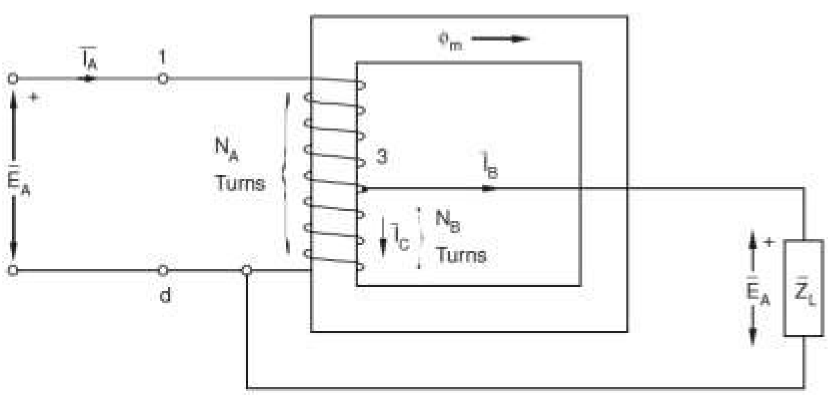


Fig: 2.26

Neglecting the magnetizing ampere-turns needed by the core for producing flux, as in an ideal transformer, the current I_A flows through only $(N_A - N_B)$ turns. If the load current is I_B, as shown by

Kirchhoff's current law, the current I_C flowing from terminal 3 to terminal 2 is (I_A - I_B). This current flows through N_B turns. So, the requirement of a net value of zero ampere-turns across the core demands that

$$(N_A - N_B)\,\bar{I}_A + (\bar{I}_A - \bar{I}_B)\,N_B = 0$$

or $$N_A\,\bar{I}_A - N_B\,\bar{I}_B = 0$$

Hence, just as in a two-winding transformer,

$$\frac{\bar{I}_A}{\bar{I}_B} = \frac{N_B}{N_A}$$

Consequently, as far as voltage, current converting properties are concerned, the autotransformer of Figure: 26 behaves just like a two-winding transformer. However, in the autotransformer we don't need two separate coils, each designed to carry full load values of current.

Parallel Operation of Transformers

It is economical to install numbers of smaller rated transformers in parallel than installing a bigger rated electrical power transformers. This has mainly the following advantages,

To maximize electrical power system efficiency: Generally electrical power transformer gives the maximum efficiency at full load. If we run numbers of transformers in parallel, we can switch on only those transformers which will give the total demand by running nearer to its full load rating for that time. When load increases, we can switch none by one other transformer connected in parallel to fulfil the total demand. In this way we can run the system with maximum efficiency.

To maximize electrical power system availability: If numbers of transformers run in parallel, we can shut down any one of them for maintenance purpose. Other parallel transformers in system will serve the load without total interruption of power.

To maximize power system reliability: if any one of the transformers run in parallel, is tripped due to fault of other parallel transformers is the system will share the load, hence power supply may not be interrupted if the shared loads do not make other transformers over loaded.

To maximize electrical power system flexibility: There is always a chance of increasing or decreasing future demand of power system. If it is predicted that power demand will be increased in future, there must be a provision of connecting transformers in system in parallel to fulfil the extra demand because, it is not economical from business point of view to install a bigger rated single transformer by forecasting the increased future demand as it is unnecessary investment of money. Again if future demand is decreased, transformers running in parallel can be removed from system to balance the capital investment and its return.

Conditions for Parallel Operation of Transformers

When two or more transformers run in parallel, they must satisfy the following conditions for satisfactory performance. These are the conditions for parallel operation of transformers.

- ***Same voltage ratio of transformer.***
- ***Same percentage impedance.***
- ***Same polarity.***
- ***Same phase sequence.***
- ***Same Voltage Ratio***

Same voltage ratio of transformer.

If two transformers of different voltage ratio are connected in parallel with same primary supply voltage, there will be a difference in secondary voltages. Now say the secondary of these transformers are connected to same bus, there will be a circulating current between secondaries and therefore between primaries also. As the internal impedance of transformer is small, a small voltage difference may cause sufficiently high circulating current causing unnecessary extra I2R loss.

Same Percentage Impedance

The current shared by two transformers running in parallel should be proportional to their MVA ratings. Again, current carried by these transformers are inversely proportional to their internal impedance. From these two statements it can be said that, impedance of transformers running in parallel are inversely proportional to their MVA ratings. In other words, percentage impedance or per unit values of impedance should be identical for all the transformers that run in parallel.

Same Polarity

Polarity of all transformers that run in parallel, should be the same otherwise huge circulating current that flows in the transformer but no load will be fed from these transformers. Polarity of transformer means the instantaneous direction of induced emf in secondary. If the instantaneous directions of induced secondary emf in two transformers are opposite to each other when same input power is fed to both of the transformers, the transformers are said to be in opposite polarity. If the instantaneous directions of induced secondary e.m.f in two transformers are same when same input power is fed to the both of the transformers, the transformers are said to be in same polarity.

Same Phase Sequence

The phase sequence or the order in which the phases reach their maximum positive voltage, must be identical for two parallel transformers. Otherwise, during the cycle, each pair of phases will be short circuited.

The above said conditions must be strictly followed for parallel operation of transformers but totally identical percentage impedance of two different transformers is difficult to achieve practically, that is why the transformers run in parallel may not have exactly same percentage impedance but the values would be as nearer as possible.

Why Transformer Rating in kVA?

An important factor in the design and operation of electrical machines is the relation between the life of the insulation and operating temperature of the machine. Therefore, temperature rise resulting from the losses is a determining factor in the rating of a machine. We know that copper loss in a transformer depends on current and iron loss depends on voltage. Therefore, the total loss in a transformer depends on the volt-ampere product only and not on the phase angle between voltage and current i.e., it is independent of load power factor. For this reason, the rating of a transformer is in kVA and not kW.

Acknowledgement

The committee members gratefully acknowledge google, scribd, NPTEL, openoffice, sumantra pdf, scilab for myriad suggestions and help for preparing this lecture note. The committee members also wants to express their gratitude to the persons out there who thinks knowledge should be free and be accessible and sharable without any restrictions so that every single person on this planet has the same opportunity to explore, expand and become enlightened by the collective gifts of mankind.

However apart from this lecture note students/readers are strongly recommended to follow the below mentioned books in the references and above all confer with the faculty for thorough knowledge of this authoritative subject of electrical engineering.

References

[1] P.S. Bimbhra, Electrical Machinery, Khanna Pub, 7th Ed.

[2] P.S. Bimbhra, Generalized Theory of Electrical Machine, Khanna Pub., 2nd Ed.

[3] M.G. Say, The Performance and Design of Alternating Current Machines, CBS Pub., 3rd Ed.

[4] A.E.Fitzgerald, S.D.Umans, C.Kingsley Jr., Electric Machinery, McGraw Hill Education(india) pvt. Ltd., 6th Ed.

[5] A.S. langsdorf, Theory of Alternating Current Machinery, McGraw Hill Education(india) pvt. Ltd., 2nd Ed.

[6]T.C. Lloyd, A.F. Puchstein and A.G. Conrad, Alternating-Current Machines, John Wiley& Sons pub., 3rd Ed.

[7] Irving Kosow, Electric Machinery and Transformers, Pearson India, 2nd Ed.

[8] C.L.Dawes, A course in Electrical Engineering, Vol-I, Vol-II, Mcgraw Hill Book Company Inc.

[9] P.K. Mukherjee & S. Chakravorty, Electrical Machines, Dhanpat Rai Publications, 2nd Ed.

[10] I.J Nagrath & D.P.Kothari, Electrical Machines, McGraw Hill Education(india) pvt. Ltd., 4th Ed.

[11] N.N.Parker Smith, Problems in Electrical Engineering, CBS Pub. 9th Ed.

Best of Luck to All the Students

Principle of Operation

DC motor operates on the principle that when a current carrying conductor is placed in a magnetic field, it experiences a mechanical force given by F = BIL newton. Where 'B' = flux density in wb, 'I' is the current and 'L' is the length of the conductor. The direction of force can be found by Fleming's left hand rule. From the point of construction, there is no difference between a DC generator and DC motor. Figure 3.1 shows a multipolar DC motor. Armature conductors are carrying current downwards under North Pole and upwards under South Pole. When the field coils are excited, with current carrying armature conductors, a force is experienced by each armature conductor whose direction can be found by Fleming's left hand rule. This is shown by arrows on top of the conductors. The collective force produces a driving torque which sets the armature into rotation. The function of a commutator in DC motor is to provide a continuous and unidirectional torque.

In DC generator the work done in overcoming the magnetic drag is converted into electrical energy. Conversion of energy from electrical form to mechanical form by a DC motor takes place by the work done in overcoming the opposition which is called the 'back emf'.

Back EMF

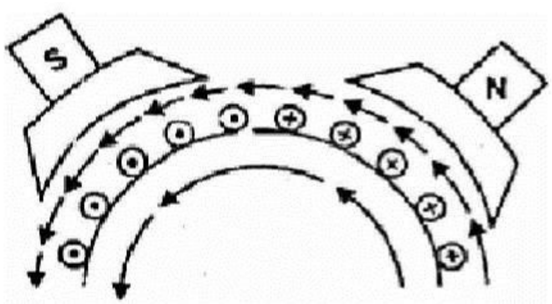


Fig. 3.1 Generation of force in DC motor

It is the dynamically induced emf in the armature conductors when the armature rotates following principle of DC motor. The direction of this induced emf can be determined using Fleming’s right hand rule. This emf act in opposition to the supply voltage of the armature. It opposes the supply voltage that is why it is called back emf. The value of this induced emf is same as the value of the emf induced in dc generator. The work done in overcoming this opposition is converted into mechanical energy.

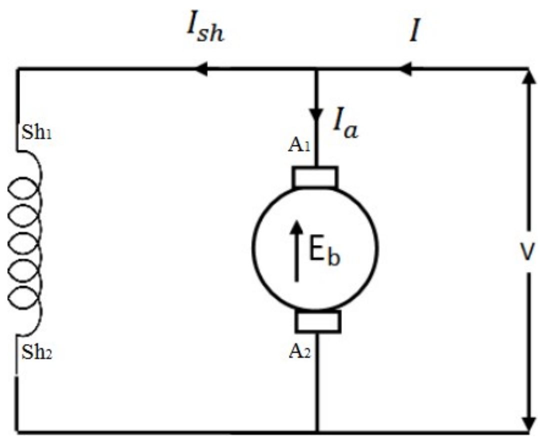


Fig. 3.2 Schematic diagram of DC shunt motor

Fig shown 3.2 a DC shunt motor the rotating armature generating the back emf E_b. The armature current can be written as

$$I_a = \frac{V_T - E_b}{r_a}$$

Where r_a is armature resistance,

$$\boxed{E_b = \frac{P\phi ZN}{60A}}$$

Armature current is proportional to back emf. So back emf is a controlling factor of armature current.

Torque Equation

Let T_a = armature torque in N –m developed by the armature of a motor running at N.rps.

Therefore $P = T_a \text{ X } \frac{2\pi N}{60}$ Watts

Electrical equivalent of mechanical power developed $P_m = E_b I_a$

$$P_m = E_b I_a = P = T_a \times \frac{2\pi N}{60}$$

$T_a = \frac{E_b I_a}{2\pi N} \times 60$ Also, on substituting for Eb i.e., $E_b = \frac{P\phi ZN}{60A}$

Therefore,

$$T_a = \frac{I_a \times 60}{2\pi N} \times \frac{P\phi ZN}{60A}$$

$$T_a = \frac{I_a}{2\pi} \times \frac{P\phi Z}{A}$$

$$\boxed{T_a = \frac{P\phi Z I_a}{2\pi A}} \text{ N-m}$$

From the above equation for torque, it is seen that

(i) $T_a = k\phi I_a$

(ii) (ii) $T_a \propto I_a^2$ - For series motor (because $\phi \propto I_a$) before saturation. After saturation $T_a \propto I_a$

(iii) $T_a \propto I_a$ - For shunt motor. (because ϕ is constant in a shunt motor)

3.4 Characteristics of DC Motors

There are three important characteristics-

1. Armature torque vs armature current T_a vs I_a *(Electrical characteristics)*
2. Speed vs armature current characteristic N vs I_a
3. Speed vs torque N vs T_a (Mechanical characteristics)

Characteristics of DC shunt motor

Armature torque vs armature current T_a vs I_a characteristics

For a shunt motor flux can be assumed practically constant (at heavy loads, decreases, due to increased armature reaction)

$T_a = k\phi I_a$

ϕ is constant, $T_a \propto I_a$

Therefore electrical characteristic of a shunt motor is a straight line through origin shown by dotted line in figure 3.3. Armature reaction weakens the flux hence T_a ys I_a characteristic bends as shown by dark line in figure 3.3, Shunt motors should never be started on heavy loads, since it draws heavy current under such condition.

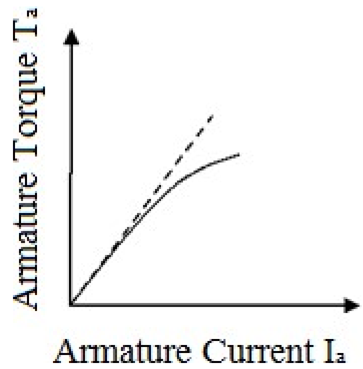


Fig. 3.3 Torque Current Characteristic of DC shunt motor

Speed vs armature current N_a vs I_a characteristics

$$N \propto \frac{E_b}{\phi}$$

$$N = \frac{V - I_a r_a}{k\phi}$$

$$N = \frac{V}{k\phi} - \frac{I_a r_a}{k\phi}$$

V is constant and in dc shunt motor ϕ is also constant. Thus with armature current speed drops and the speed current characteristics is drooping in nature is shown in figure 3.4.

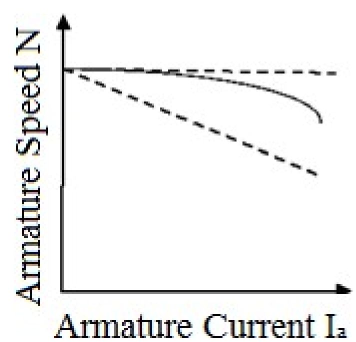


Fig. 3.4 Speed vs armature current characteristics of DC shunt motor

Speed vs armature torque N_a vs T_a characteristics

$$N = \frac{V}{k\phi} - \frac{I_a r_a}{k\phi}$$

$$I_a = \frac{T_a}{k\phi}$$

$$N = \frac{V}{k\phi} - \frac{r_a}{k\phi^2} T_a$$

Thus with increase with torque the speed of DC shunt motor decreases. The nature of the characteristics is drooping in nature shown in figure 3.5.

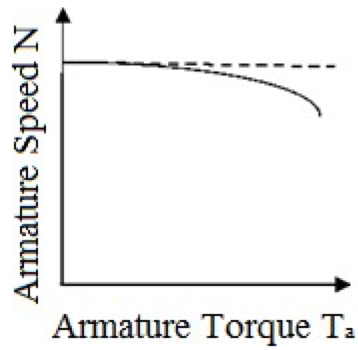


Fig. 3.5 Speed vs armature torque characteristics of DC shunt motor

Characteristics of DC series motor

Armature torque vs armature current T_a vs I_a characteristics

$T_a = k\phi I_a$

$T_a \propto I_a^2$ - For series motor (because $\phi \propto I_a$) before saturation

After saturation ϕ becomes constant thus $T_a \propto I_a$

At light loads, I_a and hence ϕ is small. But as I_a increases T_a increases as the square of the current up-to saturation. After saturation ϕ becomes constant, the characteristic becomes a straight line as shown in Figure 3.6. Therefore a series motor develops a torque proportional to the square of the armature current. This characteristic is suited where huge starting torque is required for accelerating heavy masses.

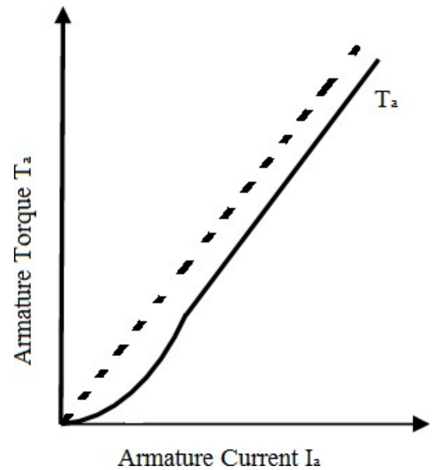


Fig. 3.6 Torque Current Characteristic of DC series motor

Speed vs armature current N_a vs I_a characteristics

$$N \propto \frac{E_b}{\phi}$$

$$N = \frac{V - I_a r_a}{k\phi}$$

$$N = \frac{V}{k\phi} - \frac{I_a r_a}{k\phi}$$

$$I_a \propto \phi$$

$$N = \frac{V}{k k_1 I_a} - \frac{k_1 r_a}{k}$$

$$N \propto \frac{1}{I_a}$$

If I_a increases, speed decreases. This characteristic is shown in figure 3.7. Therefore the speed is inversely proportional to armature current I_a. When load is heavy I_a ia heavy thus speed is low. When load is low I_a is low thus speed becomes dangerously high. Hence series motor should never started without load on it.

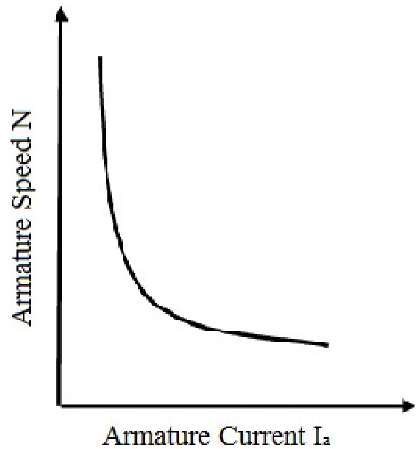


Fig. 3.7 Speed vs armature current characteristics of DC series motor

3.4.2.2 Speed vs armature torque N_a vs T_a characteristics

$$N=\frac{V}{k\phi}-\frac{I_a r_a}{k\phi} \text{ and } \phi = k_1 I_a$$

$$\therefore N=\frac{V}{kk_1 I_a}-\frac{I_a r_a}{kk_1 I_a}$$

$$\Rightarrow N=\frac{V}{kk_1 I_a}-\frac{r_a}{kk_1}$$

Now, $T_a = kI_a^2 \quad \therefore I_a = \sqrt{\frac{T_a}{k}}$

Substituting I_a

$$N=\frac{V\sqrt{k}}{kk_1\sqrt{T_a}}-\frac{r_a}{kk_1}$$

$$\Rightarrow N=\frac{\text{Const.}}{\sqrt{T_a}}-\text{Const.}$$

Thus Speed is inversely proportional to torque. The characteristics is shown in figure 3.8.

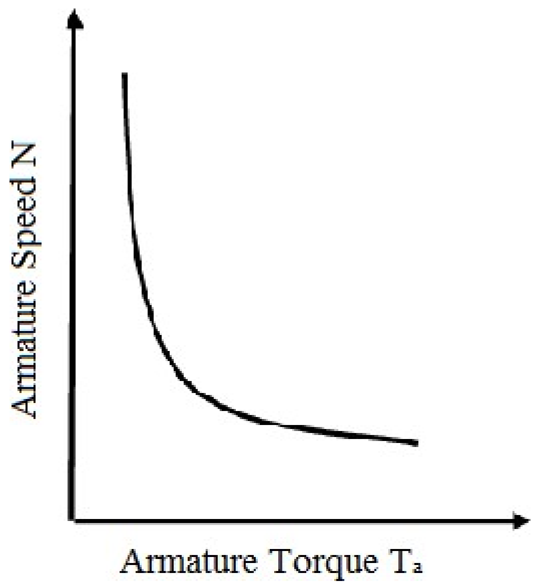


Fig. 3.8 Speed vs armature torque characteristics of DC series motor

Characteristics of DC compound motor

There are two different types of compound motors in common use, they are the cumulative compound motor and the differential compound motor. In the cumulative compound motor, the field produced by the series winding aids the field produced by the shunt winding. The speed of this motor falls more rapidly with increasing current than does that of the shunt motor because the field increases. In the differential compound motor, the flux from the series winding opposes the flux from the shunt winding. The field flux, therefore, decreases with increasing load current. Because the flux decreases, the speed may increases with increasing load. Depending on the ratio of the series-to-shunt field ampere-turns, the motor speed may increases very rapidly.

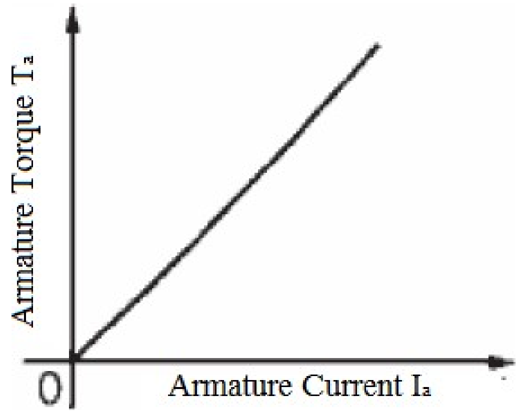


Fig. 3.9 Armature torque vs armature current characteristics of DC compound motor

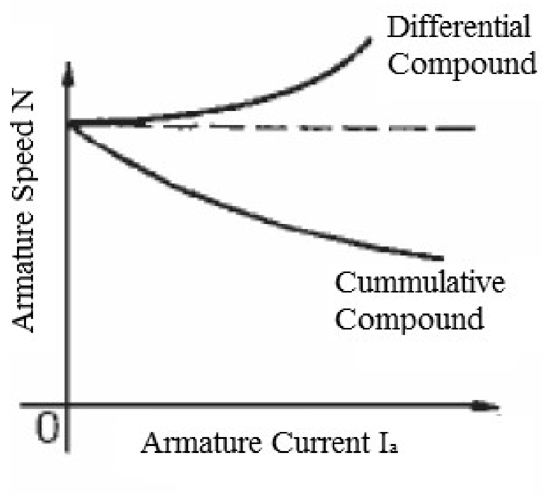


Fig. 3.10 Speed vs armature current characteristics of DC compound motors

The torque-speed (c/s) of a cumulatively compound D.C motor

In the cumulative compounded D.C. motor, there is a component of flux which is constant and another component which is proportional to its armature current (and thus to its load). Therefore, the cumulatively compounded motor has a higher starting torque than a shunt motor (whose flux is constant) but a lower starting torque than a series motor (whose entire flux is proportional to armature current). At light loads, the series field has a very small effect, so the motor behaves approximately as a shunt D.C. motor. As the load gets very large, the series flux becomes quite important and the torque-speed curve begins to look like a series motor's (c/s). A comparison of the torque-speed (c/s) of each of these type of machines is shown in figure 3.11.

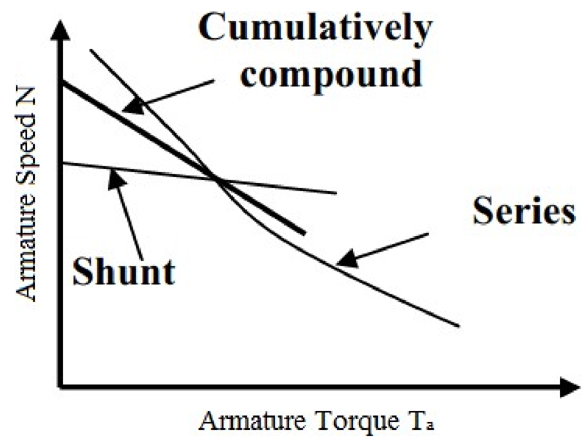


Fig. 3.11 Speed vs armature torque characteristics of DC motors

The torque-speed (c/s) of a differentially compound D.C motor

In a differentially compound D.C. motor, the shunt magneto motive force and series magneto motive force subtract from each other. This means that as the load on the motor increases, I_a increases and the flux in the motor decreases. But as the flux decreases, the speed of the motor increases. This speed increases causes anther increases in load, which further increases I_a, further decreasing the flux, and increasing the speed again. The result is that a differentially compounded motor is unstable and tends to run away. It is so bad that a differentially compounded motor is unsuitable for any application. The torque speed characteristics is shown in figure 3.12.

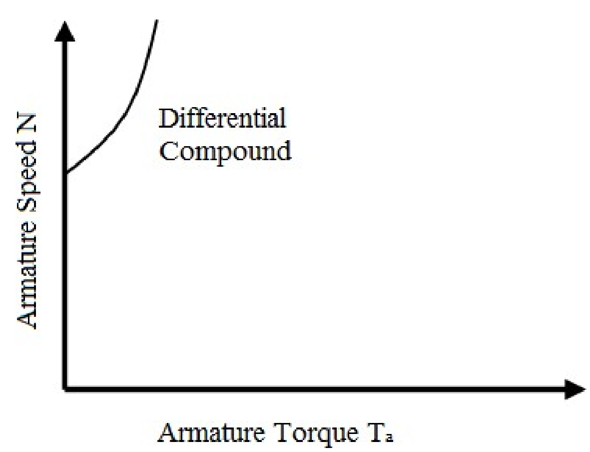


Fig. 3.12 Speed vs armature torque characteristics of DC differential compound motor

Application of DC motors

Application of DC shunt motor

The characteristics of a DC shunt motor give it a very good speed regulation, and it is classified as a constant speed motor, even though the speed does slightly decrease as load is increased. Shunt wound motors are used in industrial and automotive applications where precise control of speed and torque are required.

Application of DC series motor

For a given input current, the starting torque developed by a DC series motor is greater than that developed by a shunt motor. Hence series motors are used where huge starting torques are necessary. Ex. Cranes, hoists, electric traction etc. The DC series motor responds by decreasing its speed for the increased in load. The current drawn by the DC series motor for the given increase in load is lesser than DC shunt motor. The drop in speed with increased load is much more prominent in series motor than that in a shunt motor. Hence series motor is not suitable for applications requiring a constant speed.

Application of DC compound motor

Cumulative compound wound motors are virtually suitable for almost all applications like business machines, machine tools, agitators and mixers etc. Compound motors are used to drive loads such as shears, presses and reciprocating machines.

Differential compound motors are seldom used in practice (because of rising speed characteristics).

Armature Reaction

The action of magnetic field set up by armature current on the distribution of flux under main poles of a DC machine is called armature reaction.

When the armature of a DC machines carries current, the distributed armature winding produces its own mmf. The machine air gap is now acted upon by the resultant mmf distribution caused by the interaction of field ampere turns (AT_f) and armature ampere turns (AT_a). As a result the air gap flux density gets distorted.

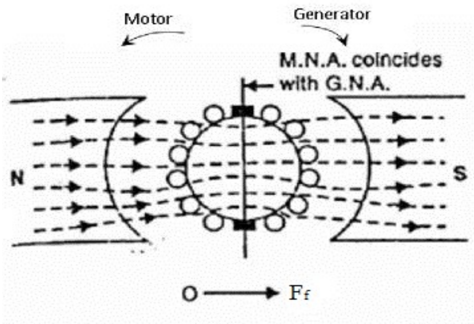


Fig. a

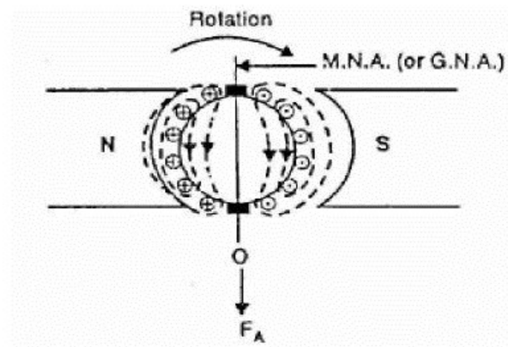


Fig. b

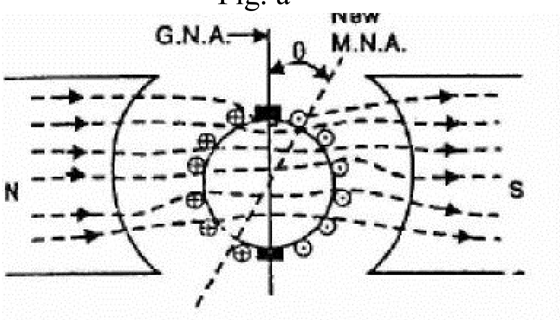


Fig. c

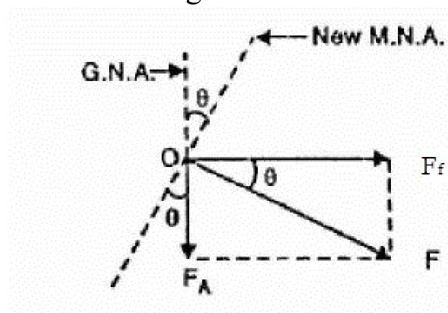


Fig. d

Figure a shows a two pole machine with single equivalent conductor in each slot and the main field mmf (F_f) acting alone. The axis of the main poles is called the direct axis (d-axis) and the interpolar axis is called quadrature axis (q-axis). It can be seen from the Figure b that armature mmf (F_a) is along the interpolar axis. F_a which is at 90^0 to the main field axis is known as cross magnetizing mmf.

Figure c shows the practical condition in which a DC machine operates when both the Field flux and armature flux are existing. Because of both fluxes are acting simultaneously, there is a shift in brush axis and crowding of flux lines at the trailing pole tip and flux lines are weakened or thinned at the leading pole tip. (The pole tip which is first met in the direction of rotation by the armature conductor is leading pole tip and the other is trailing pole tip).

If the iron in the magnetic circuit is assumed unsaturated, the net flux/pole remains unaffected by the armature reaction though the air gap flux density distribution gets distorted. If the main pole excitation is such that the iron is in the saturated region of magnetization (practical case) the increase in flux density at one end of the poles caused by armature reaction is less than the decrease at the other end, so that there is a net reduction in the flux/pole. This is called the demagnetizing effect.

Thus it can be summarized that the nature of armature reaction in a DC machine is

1. Cross magnetizing with its axis along the q-axis.
2. It causes no change in flux/pole if the iron is unsaturated but causes reduction in flux/pole in the presence of iron saturation. This is termed as demagnetizing effect. The resultant mmf ‘F’ is shown in figure d.

Commutation

The process of reversal of current in the short circuited armature coil is called ‘Commutation’. This process of reversal takes place when coil is passing through the interpolar axis (q-axis), the coil is short circuited through commutator segments and brush.

The process of commutation of coil ‘CD’ is shown Fig. 3.13. In sub figure ‘c’ coil ‘CD’ carries 20A current from left to right and is about to be short circuited in figure ‘d’ brush has moved by a small width and the brush current supplied by the coil are as shown. In figure ‘e’ coil ‘CD’ carries no current as the brush is at the middle of the short circuit period and the brush current in supplied by coil ‘AB’ and coil ‘EF’. In sub figure ‘f’ the coil ‘CD’ which was carrying current from left to right carries current from right to left. In sub fig ‘g’ spark is shown which is due to the reactance voltage. As the coil is embedded in the armature slots, which has high permeability, the coil possess appreciable amount of self inductance. The current is changed from +20 to –20. So due to self inductance and variation in the current from +20 to –20, a voltage is induced in the coil which is given by L dI/dt. This emf opposes the change in current in coil ‘CD’ thus sparking occurs.

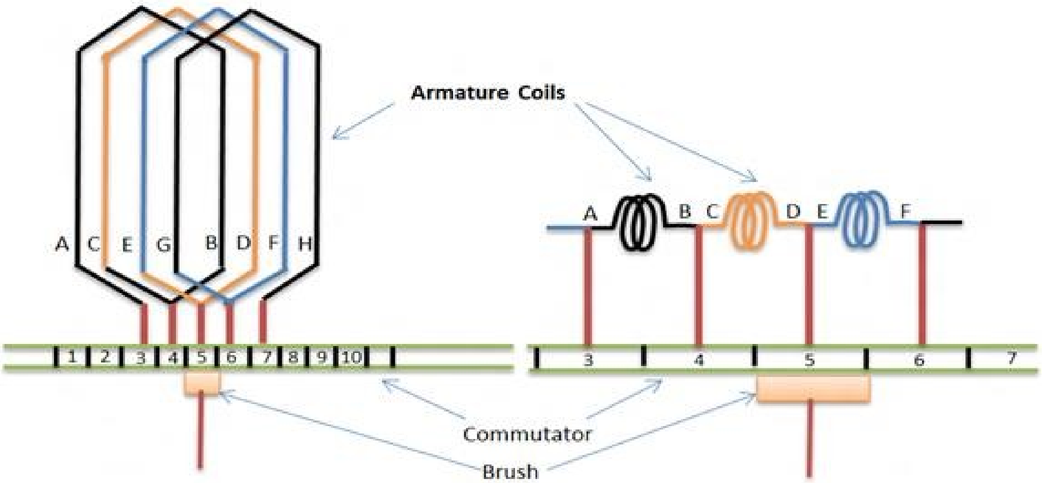


Fig a

Fig b

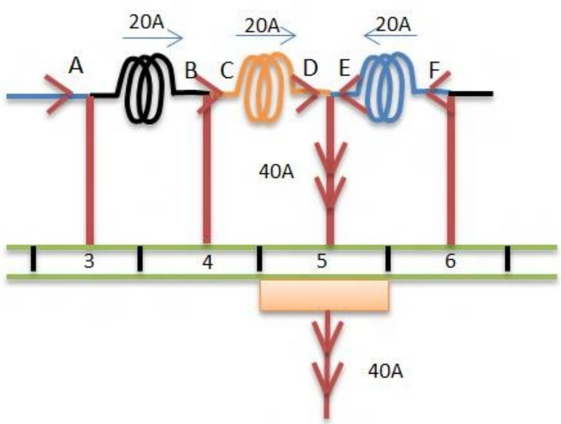


Fig.c

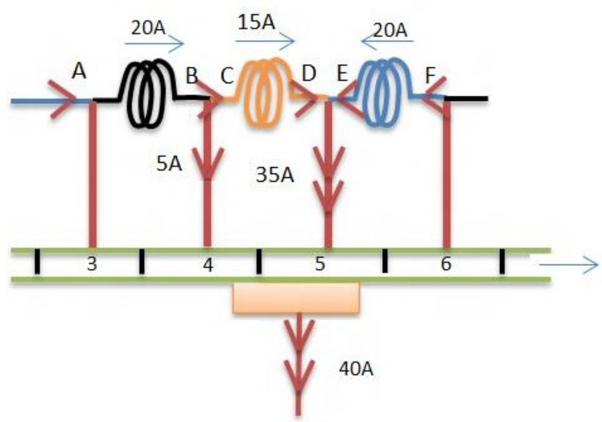


Fig.d

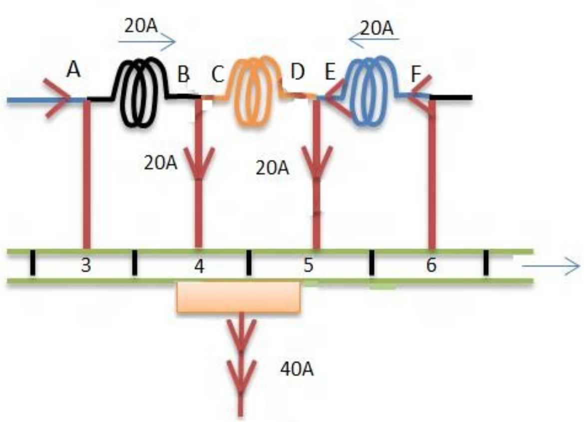


Fig.e

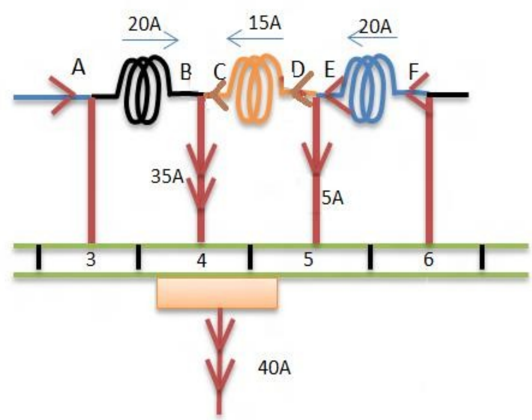


Fig.f

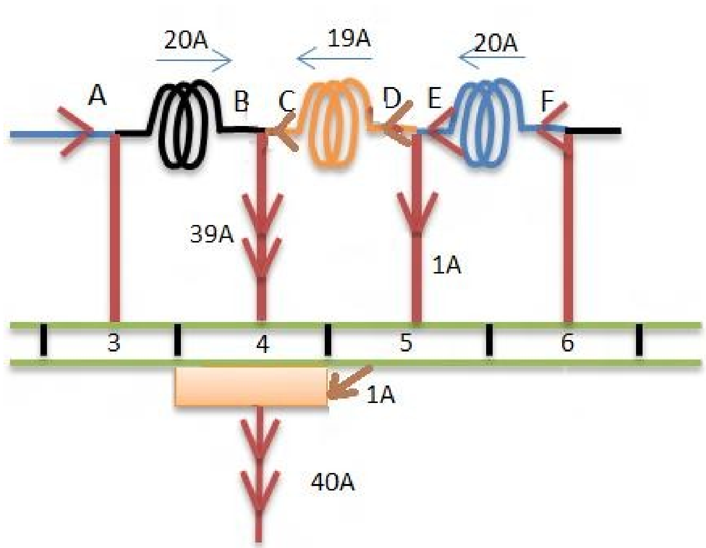


Fig.g

Fig. 3.13 a-g shows the process of commutation

Starting of DC Motor

Necessity of starter:

The current drawn by the armature is given by

$$I_a = \frac{V_T - E_b}{r_a}$$

At starting, as N=0 so Eb = 0 thus

$$I_a = \frac{V_T}{r_a}$$

Armature resistance will be very low. Therefore, the current drawn by the motor will be very high. In order to limit this high current, a starting resistance is connected in series with the armature. The starting resistance will be excluded from the circuit after the motor attains its rated speed. From there on back emf limits the current drawn by the motor.

Three Point Starter

The arrangement is shown in the figure 3.14 shows a three point starter for shunt motor.

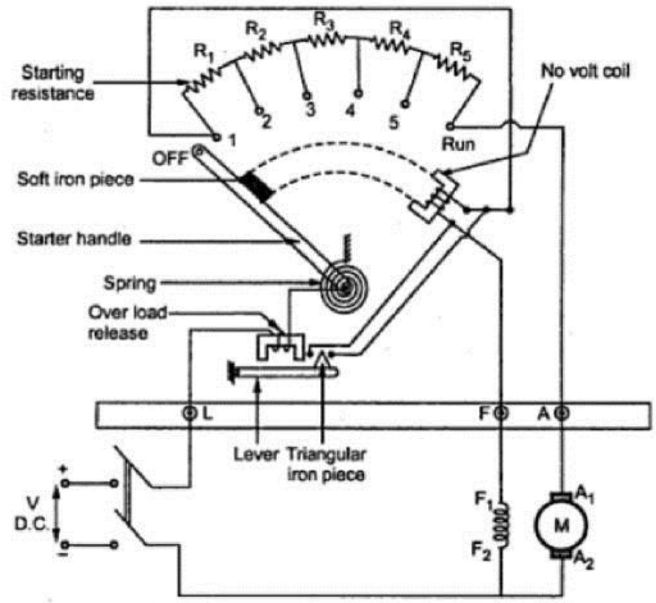


Fig. 3.14 Internal view of three point starter

It consists of resistances arranged in steps, R_1 to R_5 connected in series with the armature of the shunt motor. Field winding is connected across the supply through a protective device called 'NO – Volt Coil'. Another protection given to the motor in this starter is 'over load release coil'. To start the motor the starter handle is moved from OFF position to Run position gradually against the tension of a hinged spring. An iron piece is attached to the starter handle which is kept hold by the No-volt coil at Run position. The function of No volt coil is to get de-energized and release the handle when there is failure or disconnection or a break in the field circuit so that on restoration of supply, armature of the motor will not be connected across the lines without starter resistance. If the motor is over loaded beyond a certain predetermined value, then the electromagnet of overload release will exert a force enough to attract the lever which short circuits the electromagnet of No volt coil. Short circuiting of No volt coil results in de-energisation of it and hence the starter handle will be released and return to its off position due to the tension of the spring.

Four Point Starter

One important change is the No Volt Coil has been taken out of the shunt field and has been connected directly across the line through a Protecting resistance ‘R’. When the arm touches stud one. The current divides into three paths, 1. Through the starter resistance and the armature, 2. Through shunt field and the field rheostat and 3.Through No-volt Coil and the protecting resistance ‘R’. With this arrangement, any change of current in shunt field circuit does not affect the current passing though the NO-volt coil because, the two circuits are independent of each other. Thus the starter handle will not be released to its off position due to changes in the field current which may happen when the field resistance is varied. Fig 3.15 shows internal view of 4-point starter.

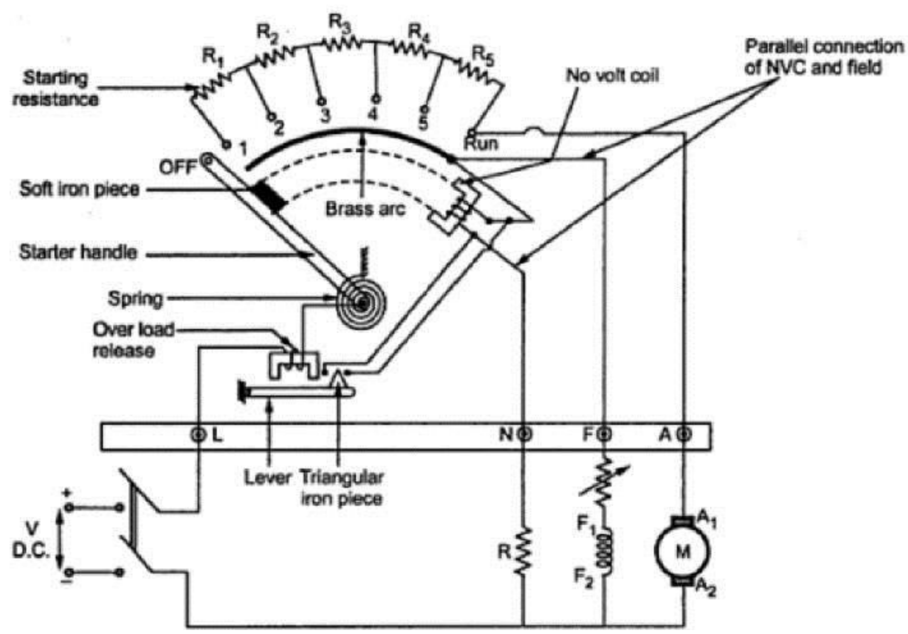


Fig. 3.15 Internal view of three point starter

DRUM CONTROLLERS

Drum controllers are used when an operator is controlling the motor directly. The drum controller is used to start, stop, reverse, and vary the speed of a motor. This type of controller is used on crane motors, elevators, machine tools, and other applications in heavy industry. As a result, the drum controller must be more rugged than the starting rheostat.

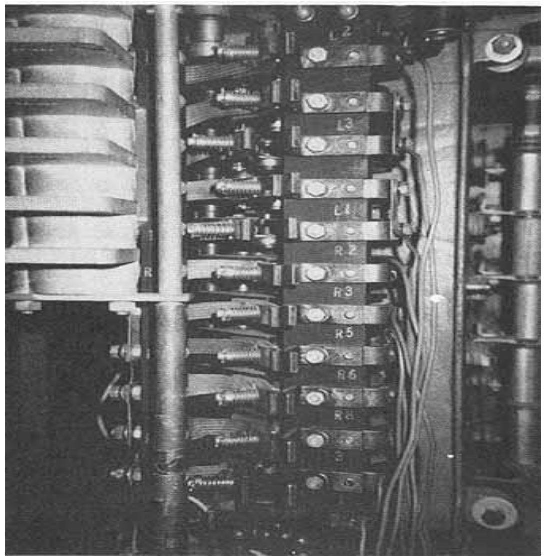


Fig. 3.16 Drum type controller shows contact fingers.

A drum controller with its cover removed is illustrated in 3.16. The switch consists of a series of contacts mounted on a movable cylinder. The contacts, which are insulated from the cylinder and from one another, are called movable contacts. There is another set of contacts, called stationary contacts, located inside the controller. These stationary contacts are arranged to touch the movable contacts as the cylinder is rotated. A handle, keyed to the shaft for the movable cylinder and contacts, is located on top of the drum controller. This handle can be moved either clockwise or counterclockwise to give a range of speed control in either direction of rotation. The handle can remain stationary in either the forward or reverse direction due to a roller and a notched wheel. A spring forces the roller into one of the notches at each successive position of the controller handle to keep the cylinder and movable contacts stationary until the handle is moved by the operator.

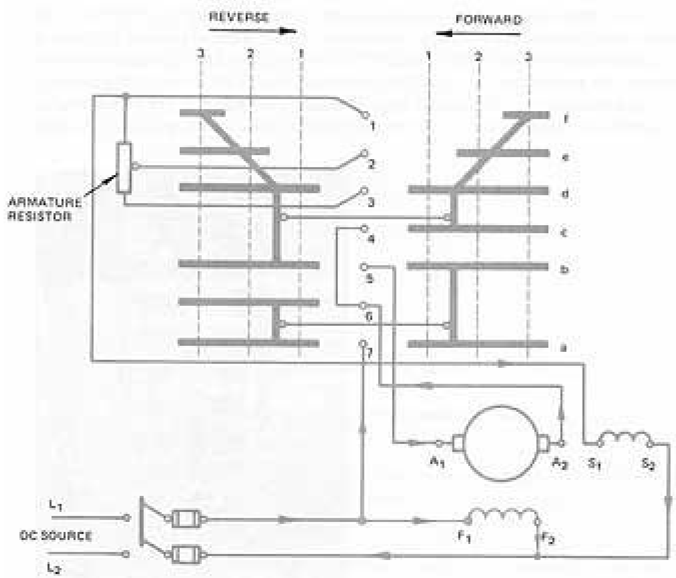


Fig. 3.17 Schematic diagram of a drum controller connected to a compound-wound motor

A drum controller with two steps of resistance is illustrated in 3.17. The contacts are represented in a flat position in this schematic diagram to make it easier to trace the circuit connections. To operate the motor in the forward direction, the set of contacts on the right must make contact with the center stationary contacts. Operation in the reverse direction requires that the set of movable contacts on the left makes contact with the center stationary contacts.

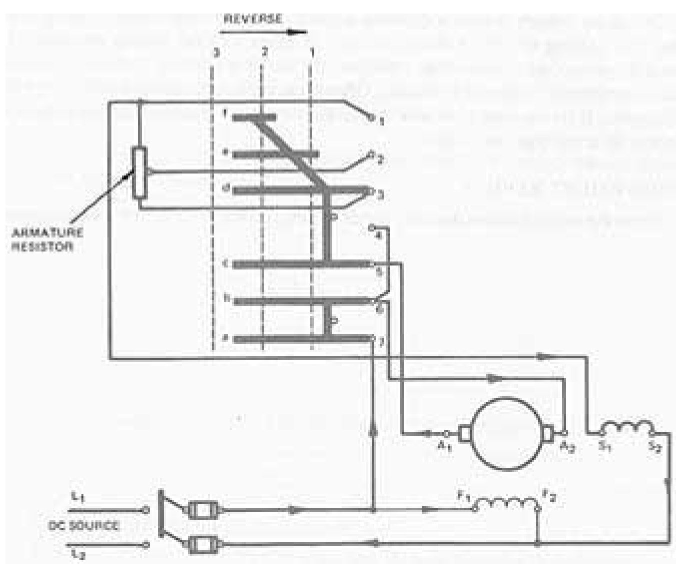


Fig. 3.18 First position of controller for reverse direction

Note in figure 3.17 that there are three forward positions and three reverse positions to which the controller handle can be set. In the first forward position, all of the resistance is in series with the armature. The circuit path for the first forward position is as follows:

1. Movable fingers a, b, c, and d contact the stationary contacts 7, 5, 4, and 3.

2. The current path is from the positive side of the line to contact 7, from 7 to a, from a to b, from b to 5, and then to armature terminal A1.

3. After passing through the armature winding to terminal A the current path is to stationary contact 6, and then to stationary contact 4.

4. From contact 4 the current path is to contact c, to d, and then to contact 3.

5. The current path then goes through the armature resistor, to the series field, and then back to the negative side of the line.

The shunt field of the compound motor is connected across the source voltage. On the second forward position of the controller handle, part of the resistance is cut out. The third forward position cuts out all of the resistance and puts the armature circuit directly across the source voltage.
In the first reverse position, all of the resistance is inserted in series with the armature. Fig. 3.18 shows the first position of the controller in the reverse direction. The current in the armature circuit's reversed. However, the current direction in the shunt and series fields is the same as the direction for the forward positions. A change in current direction in the armature only resulted in a change in the direction of rotation.

The second reverse position cuts out part of the resistance circuit. The third reverse position cuts out all of the resistance and puts the armature circuit directly across the source. Drum controllers with more positions for a greater control of speed can be obtained. However, these controllers all use the same type of circuit arrangement shown in this unit.

DC series motors require a different starting controller than shunt or compound motor. The holding circuit for the controller is in series with the starting resistance. If there is a low-voltage or no-voltage condition, the starter is returned to the off position. Drum controllers are still used frequently. Often drum controllers are used with ac as well as dc motors. It is important to be able to read the connection diagrams and the sequence diagrams on drum-type controllers.

Losses and efficiency of DC Machines

It is convenient to determine the efficiency of a rotating machine by determining the losses than by direct loading. Further it is not possible to arrange actual load for large and medium sized machines. By knowing the losses, the machine efficiency can be found by

$$\eta = \frac{\text{Output}}{\text{Output+Losses}} \text{(for Generator)}$$

$$\eta = \frac{\text{Input-Losses}}{\text{Input}} \text{(for Motor)}$$

In the process of energy conversion in rotating machines-current, flux and rotation are involved which cause losses in conductors, ferromagnetic materials and mechanical losses respectively.

Various losses occurring in a DC machine are listed below-

Total losses can be broadly divided into two types.

1) Constant losses

2) Variable losses

These losses can be further divided as

1) Constant losses –

i) Core loss or iron loss

a) Hysteresis loss

b) Eddy current loss

ii) Mechanical loss

a) Windage loss

b) Friction loss – brush friction loss and Bearing friction loss.

2) Variable losses –

i) copper loss ($I^2 r$)

a) Armature copper loss
b) Field copper loss
c) Brush contact loss

ii) Stray load loss

a) Copper stray load loss

b) Core stray load loss

Core loss or iron loss occurs in the armature core is due to the rotation of armature core in the magnetic flux produced by the field system. Iron loss consists of a) Hysteresis loss and b) Eddy current loss.

Hysteresis loss: This loss is due to the reversal of magnetization of armature core as the core passes under north and south poles alternatively. This loss depends on the volume and grade of iron, maximum value of flux density and frequency. Hysteresis loss is given by Steinmetz formula.

$W_h = K_h B_m^{1.6} f\ V$ Joule/sec or watt

Where K_h = Constant of proportionality- depends on core material.

B_m = Maximum flux density in Wb/m^2

f = Frequency in Hz

V = Volume of the armature core in m^3

Eddy Current Loss: Eddy currents are the currents set up by the induced emf in the armature core when the core cuts the magnetic flux. The loss occurring due to the flow of eddy current is known as eddy current loss. To reduce this loss the core is laminated, stacked and riveted. These laminations are insulated from each other by a thin coating of varnish. The effect of lamination is to reduce the current path because of increased resistance due to reduced cross section area of laminated core. Thus the magnitude of eddy current is reduced resulting in the reduction of eddy current loss.

Eddy Current loss is given by

$W_e = K_e B_m^2 f^2 t^2 V$ Watt

Where K_e = Constant of proportionality

B_m = Maximum flux density in Wb/m^2

f = Frequency in Hz

V = Volume of the armature core in m^3

t = Thickness of the lamination in meters

ii) Mechanical loss: these losses include losses due to windage, brush friction and bearing friction losses.

2) **Variable losses**: Variable losses consist of

(i) Copper loss:

a) Armature copper loss: This loss occurs in the armature windings because of the resistance of armature windings, when the current flows through them. The loss occurring is termed as copper loss or r loss. This loss varies with the varying load.

b) Field copper loss: This is the loss due to current flowing in the field windings of the machine. c) Brush contact drop: This is due the contact resistance between the brush and the commutator. This loss remains constant with load.

(ii) Stray load loss: The additional losses which vary with the load but cannot be related to current in a simple manner are called stray load loss. Stray load losses are.

Copper stray load loss: the loss occurring in the conductor due to skin effect and loss due to the eddy currents in the conductor set up by the flux passing through them are called copper stray load loss.

Core stray load loss: When the load current flows through the armature conductors, the flux density distribution gets distorted in the teeth and core. The flux density decreases at one end of the flux density wave and increases at the other. Since the core loss is proportional to the square of the flux density, the decrease in flux density will be less than the increase due to the increase in flux density, resulting in a net increase in the core loss predominantly in the teeth, is known as stray load loss in the core.

Further under highly saturated conditions of teeth, flux leaks through the frame and end shields causing eddy current loss in them. This loss is a component of stray load loss. Stray load loss is difficult to calculate accurately and therefore it is taken as 1 % of the output of a DC machine.

EFFICIENCY OF A DC GENERATOR:

Power flow in a DC generator is shown in figure 3.19.

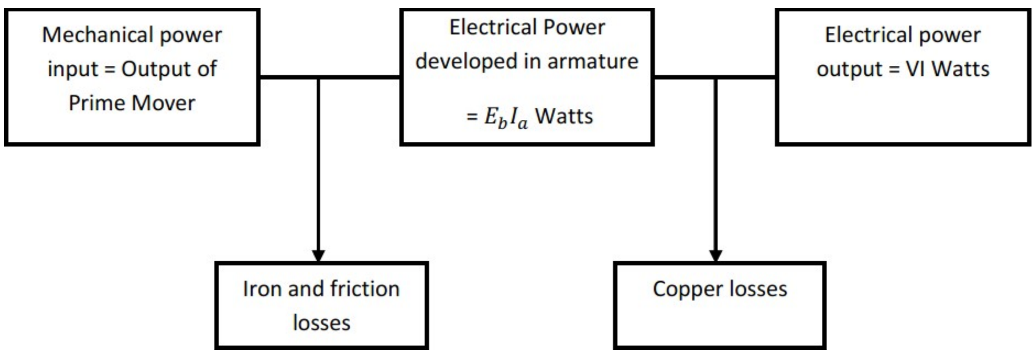


Fig. 3.19 Power flow in a DC generator

CONDITION FOR MAXIMUM EFFICIENCY

Generator output = VI;

Generator input = VI + losses.

Input $= VI + I_a^2 r_a + w_c$

If the shunt field current is negligible, then $I_a = I$

For maximum efficiency $\frac{d}{dI}(\eta) = 0$

$I_a^2 r_a = w_c$

Hence efficiency is maximum when variable loss = constant loss.

The load current corresponding to maximum efficiency is $I = \sqrt{\frac{w_c}{r_a}}$

EFFICIENCY OF DC MOTOR:

The power flow in a DC motor is shown in figure 3.20.

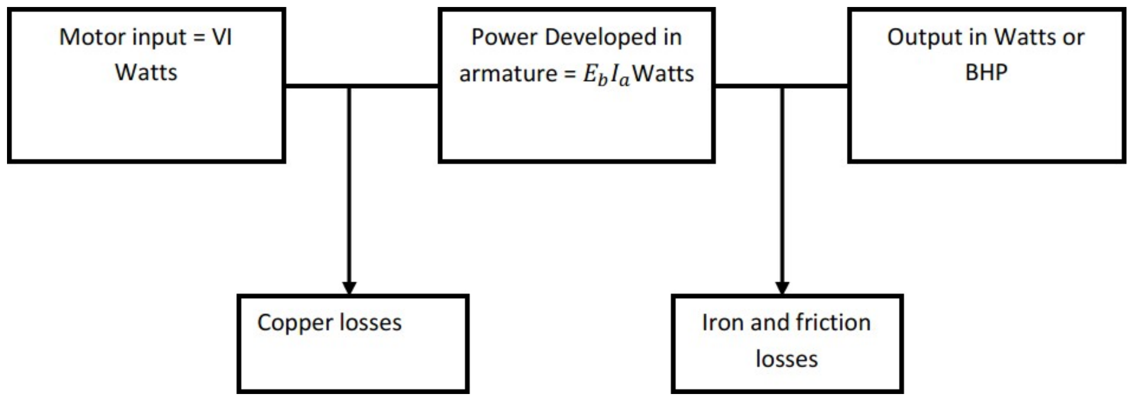


Fig. 3.20 Power flow in a DC generator

Efficiency $\eta = \dfrac{\text{Input-Losses}}{\text{Input}}$ (for Motor)

$$\eta = \frac{VI - I_a^2 r_a - w_c}{VI}$$

Efficiency is maximum when variable loss = constant loss.

Speed Control of DC Motor

DC motors are in general much more adaptable speed drives than AC motors. Speed of a DC motor can be controlled in a wide range.

$$N = \frac{E_b}{K\phi} = \frac{V - I_a r_a}{K\phi}$$

The speed equation shows that speed can be controlled by-

1. Variation of field current which varies the flux/pole (ϕ) and is known as field control.
2. Variation of armature resistance known as armature voltage control.
3. Variation of terminal voltage 'V' known as Ward Leonard method.

Field Control

For a fixed terminal voltage, $\dfrac{N_2}{N_1} = \dfrac{\phi_2}{\phi_1} = \dfrac{If_2}{If_1}$

Limitations of speed control by field control:

1. 'N' below rated speed is not possible. Because ϕ can be decreased and cannot increase.
2. $N \propto \frac{1}{\phi}$ & $T \propto \phi$ for a given armature current, this method suits for constant kW drives only where 'T' decreases if speed decreases.
3. Not suited for speed reversal.

DC Shunt Motor

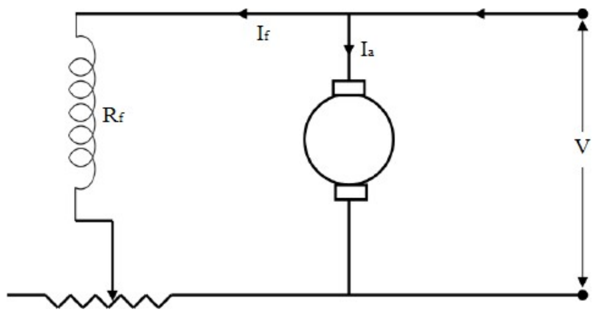


Fig. 3.21 Circuit diagram for speed control using field control method

Speed control is achieved by means of a rheostat in the field circuit as shown in the figure 3.21. The working range of the speed torque characteristics reduces with increasing speed in order for the armature current not to exceed the full load value with a weakening field.

DC Series Motor

Speed control is achieved by adjusting the field ampere turns. There are three ways for varying the field ampere turns.

A. Diverter field control

Diverter resistance R_d is connected across the field winding as shown in figure 3.22. By varying R_d the field current and hence the field ampere turns can be reduced.

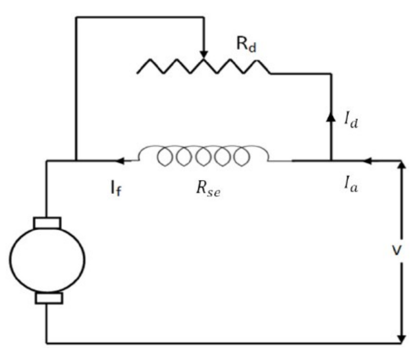


Fig. 3.22 Field diverter circuit

B. Tapped field control:

The field ampere turns are adjusted in steps by varying the number of turns included in the circuit as shown in figure 3.23. By changing number of field winding turns effective ampere turns of the field is changed thus field flux can be controlled.

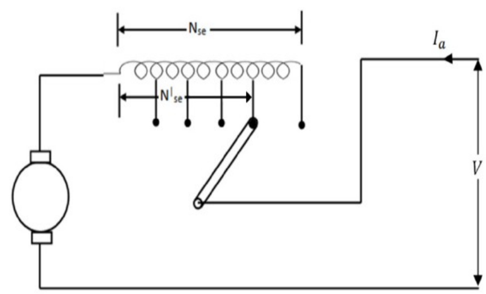


Fig. 3.23 Tapped field circuit

C. Series parallel control

In this method, the field windings are divided into two equal halves and then connected in series or parallel to control the field ampere turns as shown in figure 3.24 and 3.25respectively.

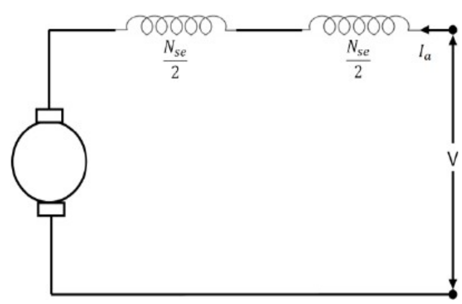


Fig. 3.24 Field circuit connected in series

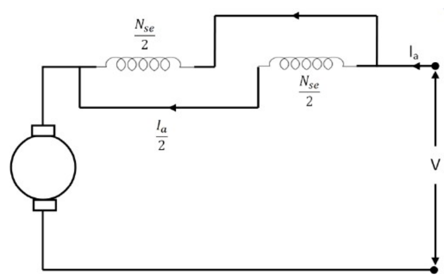


Fig. 3.25 Field circuit connected in parallel

Armature Voltage Control

In this method, applied voltage across the armature of the DC motor is varied. This method is superior to field control in the following aspects:

1. This method provides a constant torque drive.(if the ϕ and I_a are maximum, maximum torque can be obtained as $T \propto \phi I_a$)
2. Since main field ampere turns are maintained at large value, flux density distortion caused by armature reaction is limited.
3. Unlike field control scheme, speed reversal can be easily implemented.
4. This method requires a variable voltage supply which makes this method costlier.

3.12.2.1 DC Shunt Motor

Following are the armature control schemes for DC shunt motor.

1. Rheostatic control:

Here the applied armature voltage is varied by placing an adjustable resistance 'R' in series with the armature as shown in the figure 3.26. N vs T for varying 'R' is shown in figure 3.27.

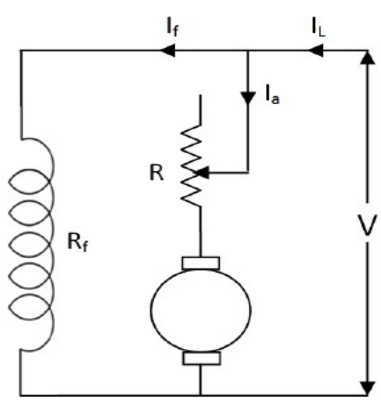


Fig. 3.27 Circuit diagram for armature resistance control

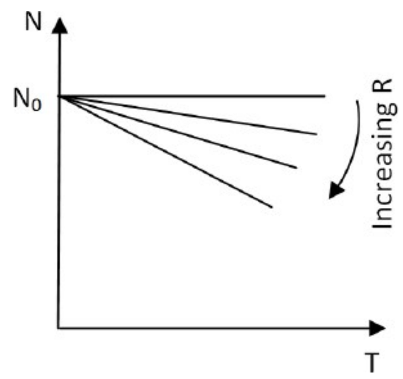


Fig. 3.28 Speed variation with torque for different resistance

Some of the limitations of the rheostatic method are:

Speeds only below rated speed

1. Range of speeds is limited because efficiency reduces drastically for large speed reductions
2. Speed regulation is poor. Because for a given resistance r_e, N varies directly with load.
3. Therefore this method is suitable for very small (fractional kW) or for short-time, intermittent slowdowns for medium sized motors.

2. Shunted armature control

In the armature rheostatic control method, the change in armature current due to change in load will affect the speed. Hence in this method the armature is shunted by an adjustable resistance as shown in figure 3.28.

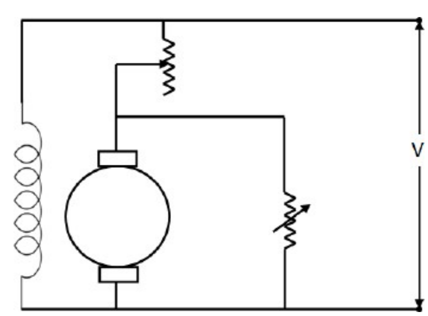


Fig. 3.28 Shunted armature control circuit

Advantages of this method are

1. Speed regulation will be better.

2. The changes in the armature current due to load will not be as effective as the armature is connected across a resistance.

Ward-Leonard Method

It is a combined armature and field control and is therefore operationally the most efficient method of speed control with a wide range. 'M_1' is the main motor whose speed control is required. The field of this motor is permanently connected across the DC supply lines. Its armature is supplied by a variable voltage derived by a Motor-Generator set. The motor M_2 act as prime mover for the generator can be AC motor or DC motor. The field of the DC generator is separately excited. The entire arrangement is shown in the figure 3.29. The reversible switch provided for the generator field makes it possible to easily reverse the generator excitation thereby reversing the voltage polarity for reversing the direction of rotation of motor. Though expensive, this arrangement can be easily adapted to feedback schemes for automatic control of speed. This method provides both constant torque and constant HP (or kW) drive.

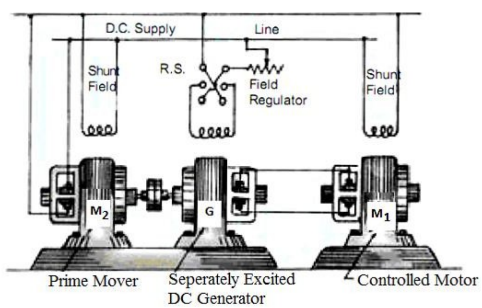


Fig. 3.29 Ward-Leonard speed control scheme

The armature and field winding of the motor are fed at maximum values at the base speed N_{base}. When armature voltage is reduced a constant torque speed control is obtained where the speed can be reduced below the base value, while the motor has full torque capability. When speed above N_{base} is required then the field is gradually weakened. The motor torque therefore reduces as its speed increases which corresponds to constant kW (or HP) drive. This is shown in the figure 3.30.

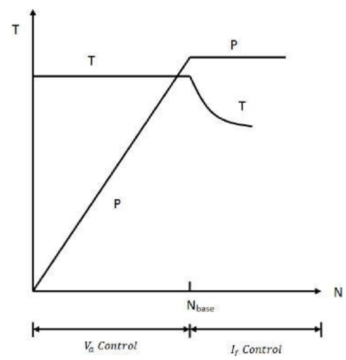


Fig. 3.30 Speed torque relation during Ward –Leonard speed control

Some of the features of the Ward Leonard system are given below:

1. As this method does not required any external resistance thus the efficiency is improved at all speeds and also when the generator emf becomes less than the back emf of the motor, the electrical power flows back from motor to generator, is converted to mechanical form and is returned to the mains via the driving AC motor.
2. Motor starts up smoothly therefore starting device is not required.
3. Reversal of speed is smoothly carried out.
4. Fine speed control from zero to rated value in both the directions are possible.

This method of speed control is used in

a. High speed elevators

b. Colliery winders

Testing of DC machines

Testing of DC machines can be broadly classified as

a) Direct method of Testing

b) Indirect method of testing

Direct method of testing

In this method, the DC machine is loaded directly by means of a brake applied to a water cooled pulley coupled to the shaft of the machine. The input and output are measured and efficiency is determined by $\eta = \frac{\text{Output}}{\text{Input}}$

It is not practically possible to arrange loads for machines of large capacity.

BRAKE TEST:
This is a direct method of testing. In this method of testing motor shaft is coupled to a Water cooled pulley which is loaded by means of weight as shown in figure 3.31.

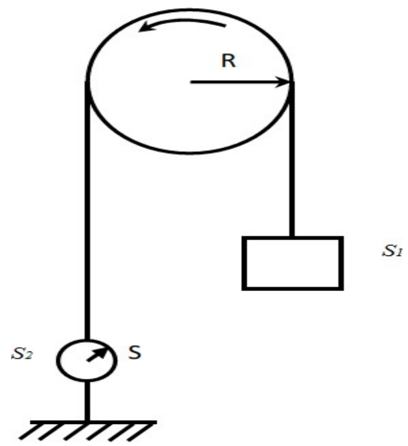


Fig. 3.31 Brake pulley arrangement for direct load test

S_1= suspended weight in kg

S_2= Reading in spring balance in kg

R = radius of pulley

n = speed in rps

V = Supply voltage

I = Full Load Current

Net pull due to friction = $(S_1 - S_2)$ kg

= 9.81 $(S_1 - S_2)$ Newton

Shaft torque = $(S_1 - S_2)$ R kg-mt

= 9.81 $(S_1 - S_2)$ R N-mt

Motor output power = T_{sh} X 2Πn Watt

= $(S_1 - S_2)$ R X 2Πn watts

Or 9.81 (S_1 — S_2)R X 2Πn watt.

Input power = VI watts

Therefore efficiency $=\eta = \dfrac{\text{Output}}{\text{Input}} = \dfrac{9.81\ (S_1 — S2)R \times 2\Pi n}{VI}$

This method of testing can be used for small motors only because for a large motor it is difficult to arrange for dissipation of heat generated at the brake.

Indirect method of testing:

In this method, the machine is not actually loaded. The losses are determined. If the losses are known, then efficiency can be determined. Swinburne's test and Hopkinson's test are commonly used on shunt motors. But, as series motor cannot be started on No-load, these tests cannot be conducted on DC series motor.

Swinburne's *Test*

For a d.c shunt motor change of speed from no load to full load is quite small. Therefore, mechanical loss can be assumed to remain same from no load to full load. Also if field current is held constant during loading, the core loss too can be assumed to remain same.

In this test, the motor is run at rated speed under *no load* condition at rated voltage. The current drawn from the supply I_{L0} and the field current I_f are recorded. Now we note that:

Input power to the motor, $P_{in} = VI_{L0}$

Cu loss in the field circuit $P_{fl} = VI_f$

Power input to the armature, $= VI_{L0} - VI_f = V(I_{L0} - I_f) = VI_{a0}$

Cu loss in the armature circuit $= I^2_{a0}\, r_a$

Gross power developed by armature $= VI_{a0} - I^2_{a0}\, r_a = \left(V - I_{a0} r_a\right) I_{a0} = E_{b0} I_{a0}$

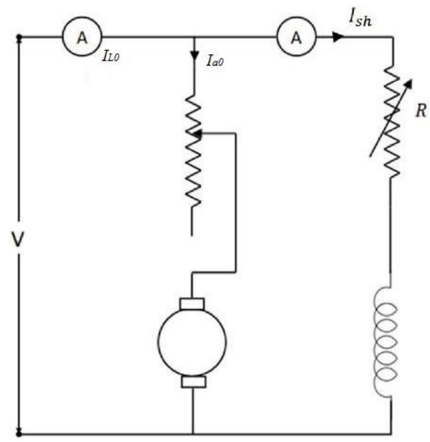


Fig. 3.32 Circuit diagram for Swinburne'*S* Test

Since the motor is operating under no load condition, net mechanical output power is zero. Hence the gross power developed by the armature must supply the core loss and friction & windage losses of the motor. Therefore,

$P_{core} + P_{friction} = (V - I_{a0} r_a) I_{a0} = E_{b0} I_{a0}$

Since, both P_{core} and $P_{friction}$ for a shunt motor remains practically constant from no load to full load, the sum of these losses is called constant rotational loss i.e.,

Constant rotational loss, $P_{rot} = P_{core} + P_{friction}$

In the Swinburne's test, the constant rotational loss comprising of core and friction loss is estimated from the above equation.

After knowing the value of P_{rot} from the Swinburne's test, we can fairly estimate the efficiency of the motor at any loading condition. Let the motor be loaded such that new current drawn from the supply is I_L and the new armature current is I_a. To estimate the efficiency of the loaded motor we proceed as follows:

Input power to the motor, $P_{in} = VI_L$

Cu loss in the field circuit $P_{fl} = VI_f$

Power input to the armature, $= VI_L - VI_f = V(I_L - I_f) = VI_a$

Cu loss in the armature circuit = $I^2_a r_a$

Gross power developed by armature = $VI_a - I^2_a r_a = (V - I_a r_a) I_a = E_b I_a$

Net mechanical output power, $P_{net\ mech} = E_b I_a - P_{rot}$

Efficiency of the loaded motor,

$$\eta = \frac{E_b I_a - P_{rot}}{VI_L} = \frac{P_{net\ mech}}{P_{in}}$$

The estimated value of P_{rot} obtained from Swinburne's test can also be used to estimate the efficiency of the shunt machine operating as a generator.

Output power of the generator, $P_{out} = VI_L$

Cu loss in the field circuit $P_{fl} = VI_f$

Output power of the armature, $= VI_L + VI_f = VI_a$

Mechanical input power, $P_{in\ mech} = VI_a + I^2_a r_a + P_{rot}$

Efficiency of the generator,

$$\eta = \frac{VI_L}{P_{in\ mech}} = \frac{VI_L}{VI_L + VI_f + I^2_a r_a + P_{rot}}$$

As this test is done at no-load condition thus the power required is very less. From the test effect of armature reaction, temperature rise, commutation etc. cannot be predicted as the machine is not actually loaded.

Load Test

To assess the rating of a machine a load test has to be conducted. When the machine is loaded, certain fraction of the input is lost inside the machine and appears as heat, increasing the temperature of the machine. If the temperature rise is excessive then it affects the insulations, ultimately leading to the breakdown of the insulation and the machine. The load test gives the information about the efficiency of a given machine at any load condition. Also, it gives the temperature rise of the machine. If the temperature rise is below the permissible value for the insulation then the machine can be safely operated at that load, else the load has to be reduced. The maximum continuous load that can be delivered by the machine without exceeding the temperature rise for the insulation used, is termed as the continuous rating of the machine. Thus the load test alone can give us the proper information of the rating and also can help in the direct measurement of the efficiency.

Hopkinson's Test

Here power drawn from the supply only corresponds to no load losses of the machines, the armature physically carries any amount of current (which can be controlled with ease). Two similar DC shunt motors are mechanically coupled. Electrically these two machines are eventually connected in parallel and controlled in such a way that one machine acts as a generator and the other as motor.

Two similar (same rating) machines are connected and coupled as shown in figure 3.33. With switch is open initially, the first machine is run as a shunt motor at rated speed. It may be noted that the second machine is operating as a separately excited generator because its field winding is excited and it is driven by the first machine. The value of the voltage across the switch is either close to twice supply voltage or small voltage. In fact the voltmeter practically reads the difference of the induced voltages in the armature of the machines. In case if the voltmeter reading is high, then the armature connection of the generator should be reversed and start afresh. Now if the voltmeter is found to read small voltage then any attempt to close the switch may result into large circulating current as the armature resistances are small. By adjusting the field current I_{fg} of the generator the voltmeter reading may be adjusted to zero ($E_g \approx E_b$) and switch is now closed. Both the machines are now connected in parallel as shown in figure 3.33.

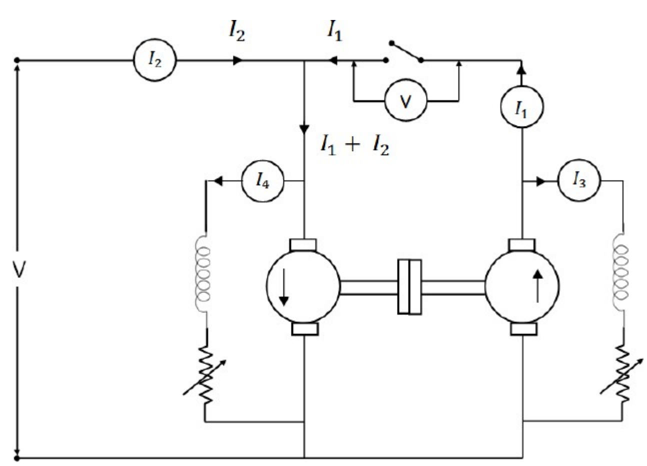


Fig. 3.33 Connection of Hopkinson's Test

After the machines are successfully connected in parallel, if the field current of generator is increased (by decreasing generator field resistance), then E_g becomes greater than E_b and both armature current of generator and motor increase, Thus by increasing field current of generator (alternatively decreasing field current of motor) one can make $E_g > E_b$ so as to make the second machine act as generator and first machine as motor. In practice, it is also required to control the field current of the motor to maintain speed constant at rated value. The interesting point to be noted here is that the armature current of generator and motor are not reflected in the supply side line. Thus current drawn from supply remains small (corresponding to losses of both the machines). The loading is sustained by the output power of the generator running the motor and vice versa. The machines can be loaded to full load current without the need of any loading arrangement.

Calculation

V= supply voltage

Motor input = $V(I_1+I_2)$

Generator output = VI_1

If it is assumed both machines have the same efficiency 'η', then,

Output of motor = η x input = η x V (I_1+I_2) = input to generator

Output of generator = η x input = η x ηV $(I_1+I_2) = \eta^2 V(I_1+I_2)$

$VI_1 = \eta^2 V(I_1+I_2)$

Therefore, $\eta = \sqrt{\dfrac{I_1}{I_1+I_2}}$

Armature copper loss in motor = $(I_1 + I_2 - I_4)^2 r_a$

Shunt field copper loss in motor = VI_4

Armature copper loss in generator = $(I_1 + I_3)^2 r_a$

Shunt field copper loss in generator = VI_3

Power drawn from supply = VI_2

Therefore stray losses = $VI_2 - [(I_1 + I_2 - I_4)^2 r_a + VI_4 + (I_1 + I_3)^2 r_a + VI_3]$ = W (say)

Stray losses/motor = $\frac{W}{2}$

<u>Therefore for generator</u>

Total losses= $(I_1+I_2)^2 r_a + VI_3 + \frac{W}{2} = W_g$

Output = VI_1, therefore

$$\eta_{generator} = \frac{VI_1}{VI_1+W_g} = \frac{output}{output+losses}$$

<u>For motor,</u>

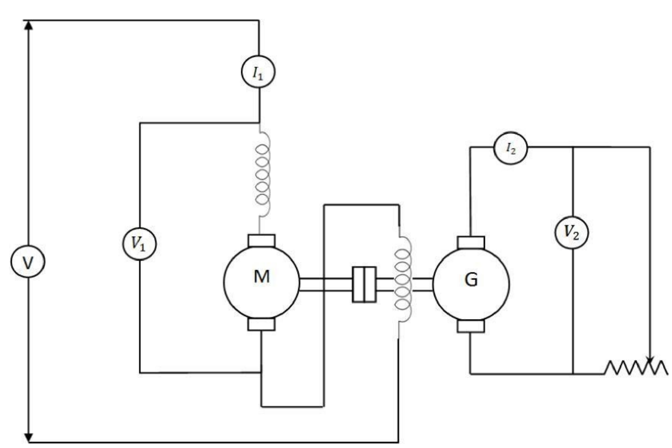
Total losses = $(I_1+I_2-I_4)^2 r_a + VI_4 + \frac{W}{2} = W_m$

Input to motor = $V(I_1+I_2)$

Therefore $\eta_{motor} = \frac{V(I_1+I_2)-W_m}{V(I_1+I_2)}$

Field's Test

Figure 3.34 shows the circuit for fields test. This test is applicable to two similar series motor. One of the machine runs as a motor and drives a generator whose output is wasted in a variable load 'R'. Both machine field coils are in series and both run at same speed so that iron and friction losses are made equal.



3.34 Circuit diagram for Field's test on DC series motor

Load resistance 'R' is varied till the motor current reaches its full load value.

V = Supply voltage

I_1 = Motor current

V_2 = Generator terminal voltage

I_2 = Load current

Input = VI_1 and output = V_2I_2

R_a and R_{se} = hot resistances.

Total losses in the set $W_t = VI_1 - V_2I_2$

Armature and Field copper losses $W_c = (R_a + 2\ r_{se})\ I_1^2 + I_a^2R_a$

Stray losses for the set = $W_t - W_c$

Stray losses per machine $W_s = \dfrac{W_t - W_c}{2}$

Motor efficiency :

Input=V_1I_1

Losses=$\left(R_a + R_{se}\right)I_1^2 + W_s = W_m$ (say)

$$\eta_{motor} = \frac{VI_1 + W_m}{VI_1}$$

Generator efficiency: η of generator is of little use, because its field winding is separately excited

Generator output = VI_2

Field copper loss = $I_1^2 r_{se}$

Armature copper loss = $I_1^2 r_a$

Total losses = $I_1^2 r_{se} + I_1^2 r_a + W_s = W_g$ (say)

$$\eta_{generator} = \frac{VI_2}{V_2 + W_g}$$

3.12.2.5Reatardation Test

This method is applicable to shunt motors and generators and is used for finding the stray losses. If armature and shunt copper losses are known for a given load, efficiency can be calculated. The circuit is shown in figure 3.35.

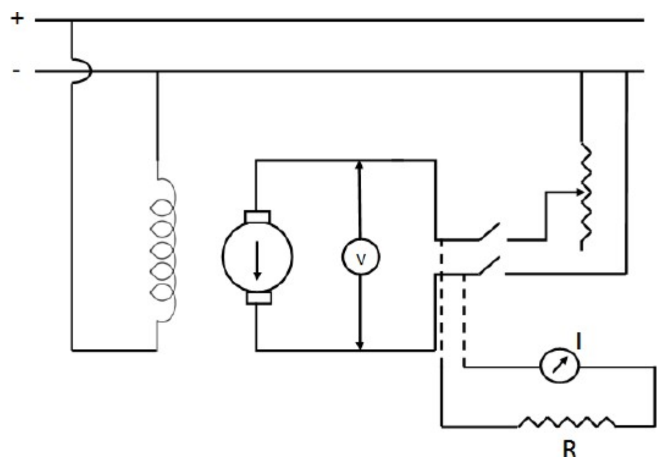


Fig. 3.35 Circuit diagram for Retardation test on DC motor

Machine is speeded up slightly beyond its rated speed and then supply is cut off from the armature while keeping the field excited. Armature will slow down and its kinetic energy is needed to meet rotational losses. i.e., friction and windage losses.

Kinetic energy of the armature = $\frac{1}{2} I\omega^2$

I = Moment of inertia of the armature

ω = Angular velocity.

Rotational losses;

N = Rate of loss of K.E.

Rate of loss of Kinetic energy W = $\frac{d}{dt}\left[\frac{1}{2} I\omega^2\right] = I\omega \frac{d\omega}{dt}$

Two quantities need to be known

(i) Moment of Inertia ‘I

(ii) $\frac{d\omega}{dt}$ or $\frac{dN}{dt}$ (because $\omega \propto N$)

(i) Finding $\frac{d\omega}{dt}$:

The voltmeter "V" in the circuit shown in Fig. 3.35 is used as per speed indicator by suitably grading it because $E \propto N$. Then the supply is cut off, the armature speed and hence voltmeter reading falls. Voltage and time at different interval are noted and a curve is drawn between time and speed as shown in fig. 3.36.

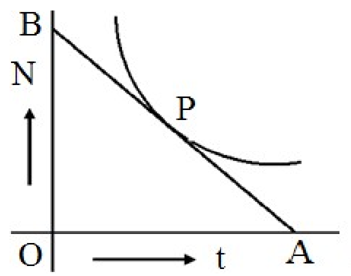


Fig. 3.36 Change of speed with time

In the fig. 3.36 AB- tangent drawn at P

Therefore $\frac{dN}{dt} = \frac{OB(rpm)}{OA(sec)}$

$$W = I \times \omega \times \frac{d\omega}{dt}$$

$$\omega = \frac{2\pi N}{60}$$

$$W = I\left(\frac{2\pi N}{60}\right)\frac{d}{dt}\left(\frac{2\pi N}{60}\right)$$

$$W = \left(\frac{2\pi}{60}\right)^2 .I.N.\frac{dN}{dt}$$

(ii) Finding Moment of Inertia "I":

There are two methods of finding the moment of inertia 'I'

(a) I is calculated:

(i) Slowing down curve with armature alone is calculated.

(ii) A fly wheel is keyed to the shaft and the curve is drawn again

For any given speed, $\frac{dN}{dt}$ and $\frac{dN}{dt_1}$ are determined as before.

Therefore $W=\left(\frac{2\pi}{60}\right)^2 .I.N.\frac{dN}{dt_1}$ 1st case

$W=\left(\frac{2\pi}{60}\right)^2 \left(I+I_1\right)N.\frac{dN}{dt_2}$ 2nd Case

The two cases are equal because losses in two cases will be almost same.

$I\frac{dN}{dt}=\left(I+I_1\right)\frac{dN}{dt}.\frac{I+I1}{I}\left(\frac{dN}{dt_2}\right)=\frac{dN}{dt_1}$

$\frac{I+I1}{I}=\frac{dN}{dt_2}$

$I = I_1 \times \frac{t_2}{t_1 - t_2}$

(b) I is eliminated:

In this method, time taken to slow down is noted with armature alone and then a retarding torque is applied electrically i.e., a non-inductive resistance is connected to the armature.

The additional loss is $I_a^2\left(R_a + R\right) or VI_a$

Let W^1 be the power then

$W=\left(\frac{2\pi}{60}\right)^2 IN.\frac{dN}{dt_1}$

$W+W^1=\left(\frac{2\pi}{60}\right)^2 IN.\frac{dN}{dt_2}$

$\frac{dN}{dt_1}$ = rate of change of speed without electrical load

$\frac{dN}{dt_2}$ = rate of change of speed with electrical load

$\frac{W+W^1}{W}=\frac{\frac{dN}{dt_2}}{\frac{dN}{dt_1}}$

or, $W=W^1 \times \frac{dt_2}{dt_1 - dt_2}$

or $W=W^1 \times \frac{t_2}{t_1 - t_2}$

Module IV

[THREE PHASE TRANSFORMER]

TOPICS

Three Phase Transformers: Constructional features of three phase transformers – three phase connection of transformers (Dd0, Dd6, Yy0, Yy6, Dy1, Dy11, Yd1, Yd11, zigzag), Scott connection, open delta connection, three phase to six phase connection, oscillating neutral, tertiary winding, three winding transformer, equal and unequal turns ratio, parallel operation, load sharing. Distribution transformers, all day efficiency, Autotransformers, saving of copper, applications, tap- changing transformers, cooling of transformers.

[Topics are arranged as per above sequence]

Three Phase Transformers

Introduction

Electric power is generated in generating stations, using three phase alternators at 11 KV. This voltage is further stepped up to 66 KV, 110 KV, 230 KV or 400 KV using 3 phase power transformers and power is transmitted at this high voltage through transmission lines. At the receiving substations, these high voltages are stepped down by 3 phase transformers to 11 KV. This is further stepped down to 400 volts at load centers by means of distribution transformers. For generation, transmission and distribution, 3 phase system is economical. Therefore 3 phase transformers are very essential for the above purpose. The sectional view of a 3 phase power transformer is shown in Fig.4.1.

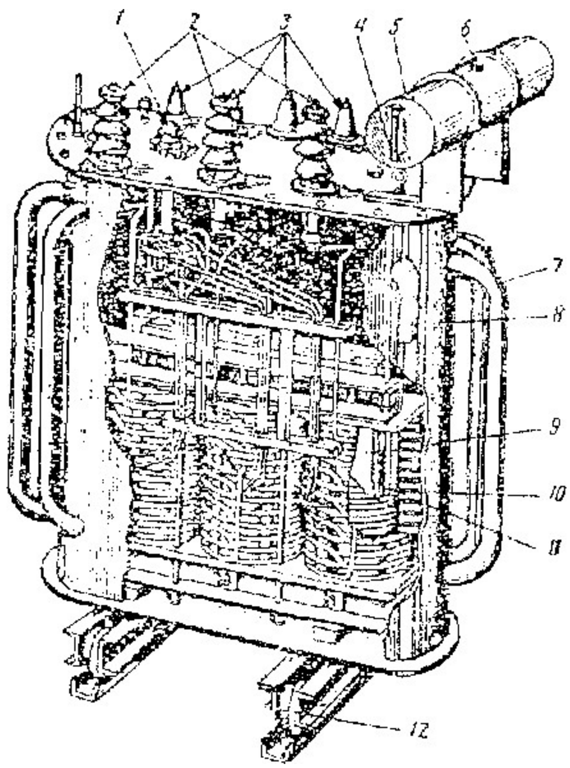


Fig. 4.1 100 KVA oil immersed power transformer

1. Tap-changer switch handle
2. Porcelain-bushing insulator (For high voltage)
3. Bushing insulators (For low voltages)
4. Oil gauge

5. Oil tank
6. Breather plug
7. Cooling pipes
8. Tank front wall
9. Core,
10. High voltage winding
11. Low voltage winding
12. Wheels or rollers.

Construction of Three phase Transformer

Three phase transformers comprise of three primary and three secondary windings. They are wound over the laminated core as we have seen in single phase transformers. Three phase transformers are also of core type or shell type as in single phase transformers. The basic principle of a three phase transformer is illustrated in fig 4.2 in which the primary windings and secondary windings of three phases are shown. The primary windings can be inter connected in star or delta and put across three phase supply.

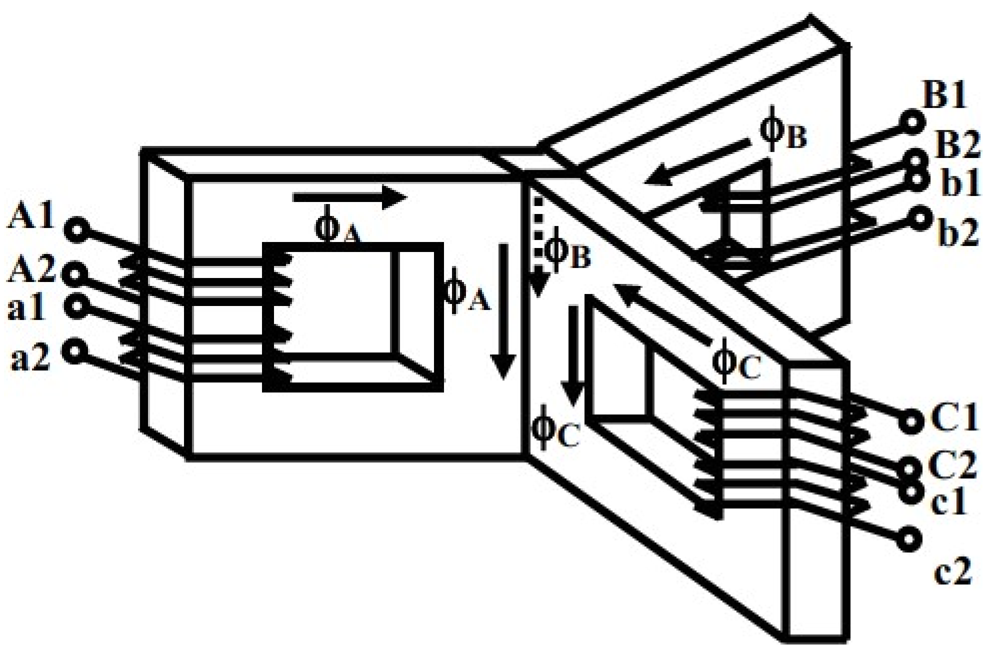


Fig. 4.2 3-phase core-type Transformer

The three cores are 120° apart and their unwound limbs are shown in contact with each other. The center core formed by these three limbs, carries the flux produced by the three phase currents I_R, I_Y and I_B. As at any instant $I_R+I_y+I_B=0$, the sum of three fluxes (flux in the center limb) is also zero.

Therefore it will make no difference if the common limb is removed. All the three limbs are placed in one plane in case of a practical transformer as shown in fig 4.3.

The core type transformers are usually wound with circular cylindrical coils. The construction and assembly of laminations and yoke of a three phase core type transformer is shown in fig 4.4 one method of arrangement of windings in a three phase transformer is shown.

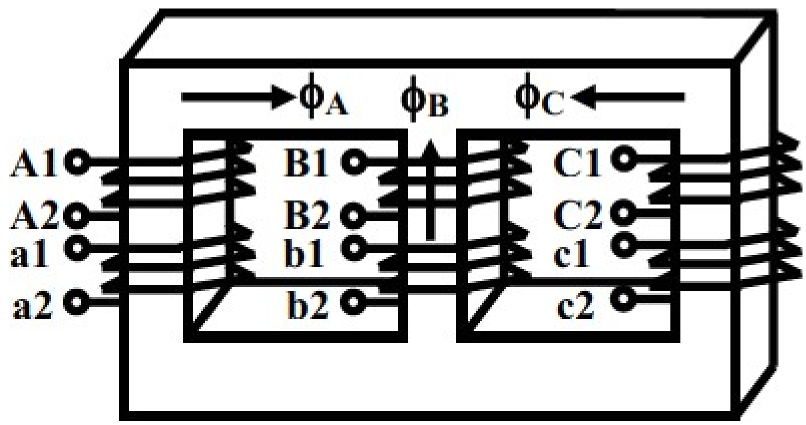


Fig. 4.3 A practical core type three phase transformer

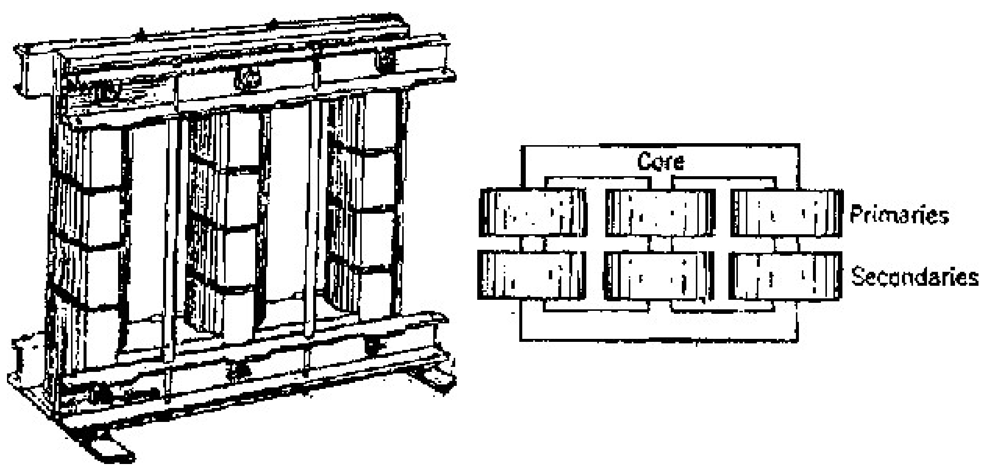


Fig. 4.4 Core type transformer windings and construction

In the other method the primary and secondary windings are wound one over the other in each limb. The low-tension windings are wound directly over the core but are, of course, insulated for it. The high tension windings are wound over the low— tension windings and adequate insulation is provided between the two windings.

The primary and secondary windings of the three phase transformer can also be interconnected as star or delta.

Three Phase Transformer connections:-

The identical single phase transformers can be suitably inter-connected and used instead of a single unit 3—phase transformer. The single unit 3 phase transformer is housed in a single tank. But the transformer bank is made up of three separate single phase transformers each with its own, tanks and bushings. This method is preferred in mines and high altitude power stations because transportation becomes easier. Bank method is adopted also when the voltage involved is high because it is easier to provide proper insulation in each single phase transformer.

As compared to a bank of single phase transformers, the main advantages of a single unit 3-phase transformer are that it occupies less floor space for equal rating, less weight costs about 20% less and further that only one unit is to be handled and connected.

There are various methods available for transforming 3 phase voltages to higher or lower 3 phase voltages. The most common connections are (i) star — star (ii) Delta—Delta (iii) Star —Delta (iv) Delta — Star.

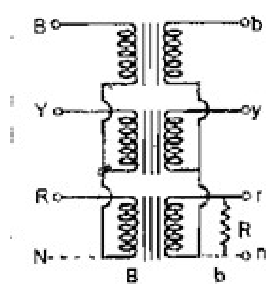


Fig 4.5 Star-star connection

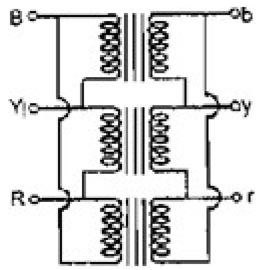


Fig. 4.6 Delta-delta connection

The star-star connection is most economical for small, high voltage transformers because the number of turns per phase and the amount of insulation required is minimum (as phase voltage is only 1/3 of line voltage. In fig. 4.5 a bank of three transformers connected in star on both the primary and the secondary sides is shown. The ratio of line voltages on the primary to the secondary sides is the same as a transformation ratio of single phase transformer.

The delta— delta connection is economical for large capacity, low voltage transformers in which insulation problem is not a serious one. The transformer connection are as shown in fig. 4.6.

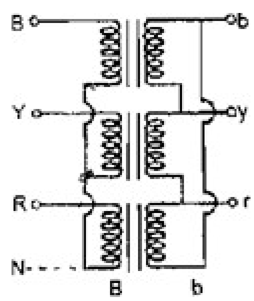


Fig. 4.7 Star-delta connection

The main use of star-delta connection is at the substation end of the transmission line where the voltage is to be stepped down. The primary winding is star connected with grounded neutral as shown in Fig. 4.7. The ratio between the secondary and primary line voltage is 1/3 times the transformation ratio of each single phase transformer. There is a 30° shift between the primary and secondary line voltages which means that a star-delta transformer bank cannot be paralleled with either a star-star or a delta-delta bank.

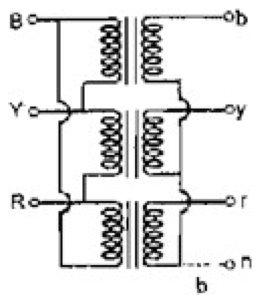


Fig. 4.8 Delta-star connection

Delta-Star connection is generally employed where it is necessary to step up the voltage. The connection is shown in fig. 4.8. The neutral of the secondary is grounded for providing 3-phase, 4-wire service. The connection is very popular because it can be used to serve both the 3-phase power equipment and single phase lighting circuits.

Vector Group of 3-phase transformer

The secondary voltages of a 3-phase transformer may undergo a *phase shift* of either +30° leading or -30° lagging or 0° i.e, no phase shift or 180° reversal with respective line or phase to neutral voltages. On the name plate of a three phase transformer, the vector group is mentioned. Typical representation of the vector group could be Ydl or Dy 11 etc. The first capital latter Y indicates that the primary is connected in star and the second lower case latter d indicates delta connection of the secondary side. The third numerical figure conveys the angle of phase shift based on *clock convention.* The minute hand is used to represent the primary phase to neutral voltage and always shown to occupy the position 12. The hour hand represents the secondary phase to neutral voltage and may, depending upon phase shift, occupy position other than 12 as shown in the figure 4.9. The angle between two consecutive numbers on the clock is 30°.

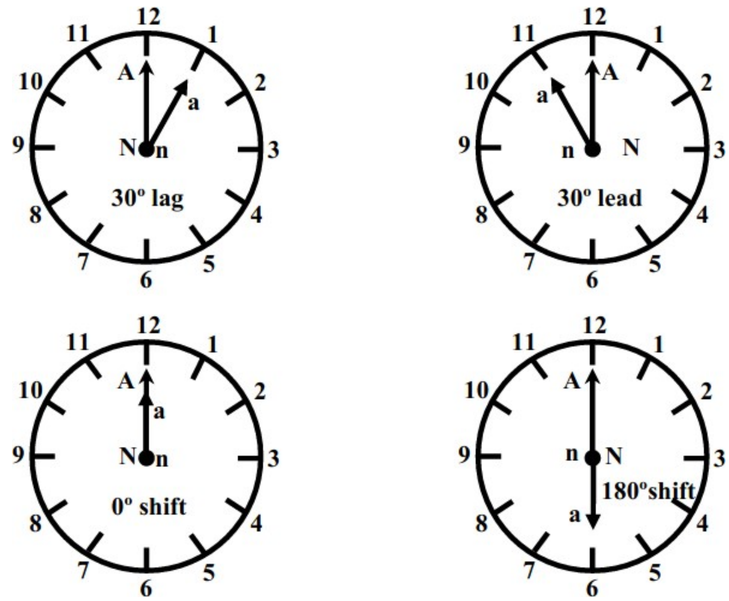


Fig. 4.9 Clock convention representing vector groups

Delta/delta (Dd0, Dd6) connection

The connection of Dd0 is shown in fig. 4.10 and the voltages on primary and secondary sides is also shown on the phasor diagram. The phase angle difference between the phase voltage of high voltage side and low voltage side is zero degree (0°).

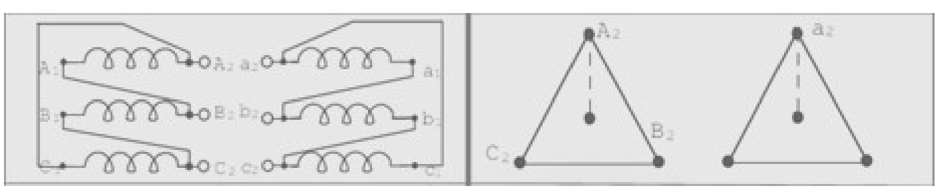


Fig 4.10 Dd0 connection and phasor diagram

The connection of Dd6 is shown in fig. 4.11 and the voltages on primary and secondary sides is also shown on the phasor diagram. The phase angle difference between the phase voltage of high voltage side and low voltage side is 180°.

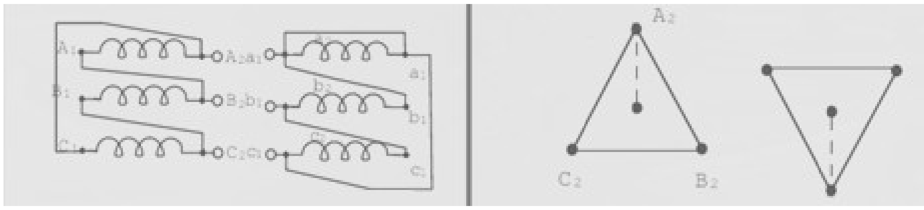


Fig 4.11 Dd6 connection and phasor diagram

This connection proves to be economical for large low voltage transformers as it increases number of turns per phase. Primary side line voltage is equal to secondary side line voltage. Primary side phase voltage is equal to secondary side phase voltage. There is no phase shift between primary and secondary voltages for Dd0 connection. There is 180° phase shift between primary and secondary voltages for Dd6 connection.

Advantages

- **Sinusoidal Voltage at Secondary:** In order to get secondary voltage as sinusoidal, the magnetizing current of transformer must contain a third harmonic component. The delta connection provides a closed path for circulation of third harmonic component of current. The flux remains sinusoidal which results in sinusoidal voltages.

- **Suitable for Unbalanced Load:** Even if the load is unbalanced the three phase voltages remains constant. Thus it suitable for unbalanced loading also.

- **Carry 58% Load if One Transfer is Faulty in Transformer Bank**: If there is bank of single phase transformers connected in delta-delta fashion and if one of the transformers is disabled then the supply can be continued with remaining tow transformers of course with reduced efficiency.

- **No Distortion in Secondary Voltage**: there is no any phase displacement between primary and secondary voltages. There is no distortion of flux as the third harmonic component of magnetizing current can flow in the delta connected primary windings without flowing in the line wires .there is no distortion in the secondary voltages.

- **Economical for Low Voltage:** Due to delta connection, phase voltage is same as line voltage hence winding have more number of turns. But phase current is (1/√3) times the line current. Hence the cross-section of the windings is very less. This makes the connection economical for low voltages transformers.
- **Reduce Cross section of Conductor:** The conductor is required of smaller Cross section as the phase current is 1/√3 times of the line current. It increases number of turns per phase and reduces the necessary cross sectional area of conductors thus insulation problem is not present.
- **Absent of Third Harmonic Voltage:** Due to closed delta, third harmonic voltages are absent.
- The absence of star or neutral point proves to be advantageous in some cases.

Disadvantages

- Due to the absence of neutral point it is not suitable for three phase four wire system.
- More insulation is required and the voltage appearing between windings and core will be equal to full line voltage in case of earth fault on one phase.

Application

- Suitable for large, low voltage transformers.
- This Type of Connection is normally uncommon but used in some industrial facilities to reduce impact of SLG faults on the primary system
- It is generally used in systems where it need to be carry large currents on low voltages and especially when continuity of service is to be maintained even though one of the phases develops fault.

<u>Star/star (YyO, Yy6) connection</u>

This is the most economical one for small high voltage transformers. Insulation cost is highly reduced. Neutral wire can permit mixed loading. Triplen harmonics are absent in the lines. These triplen harmonic currents cannot flow, unless there is a neutral wire. This connection produces oscillating neutral. Three phase shell type units have large triplen harmonic phase voltage. However three phase core type transformers work satisfactorily. A tertiary mesh connected winding may be required to stabilize the oscillating neutral due to third harmonics in three phase banks.

The connection of Yy0 is shown in fig. 4.12 and the voltages on primary and secondary sides is also shown on the phasor diagram. The phase angle difference between the phase voltage of high voltage side and low voltage side is zero degree (0°).

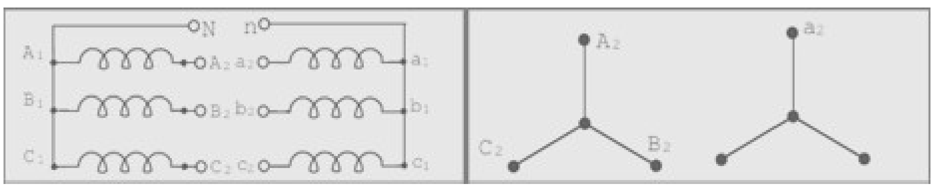


Fig .4.12 Yy0 connection and phasor diagram

The connection of Yy6 is shown in fig. 4.13 and the voltages on primary and secondary sides is also shown on the phasor diagram. The phase angle difference between the phase voltage of high voltage side and low voltage side is 180°.

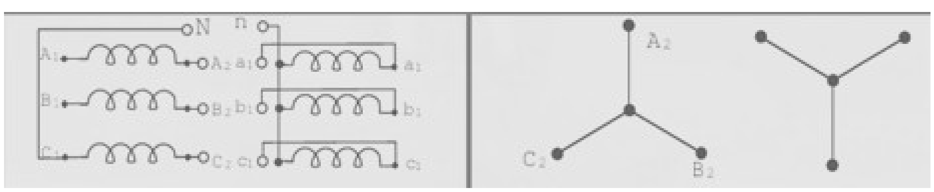


Fig 4.13. Yy6 connection and phasor diagram

- In Primary Winding Each Phase is120°electrical degrees out of phase with the other two phases.
- In Secondary Winding Each Phase is120°electrical degrees out of phase with the other two phases.

- Each primary winding is magnetically linked to one secondary winding through a common core leg. Sets of windings that are magnetically linked are drawn parallel to each other in the vector diagram. In the Y-Y connection, each primary and secondary winding is connected to a neutral point.
- The neutral point may or may not be brought out to an external physical connection and the neutral may or may not be grounded.

Advantages of Y-y connection

- **No Phase Displacement:** The primary and secondary circuits are in phase; i.e., there are no phase angle displacements introduced by the Y-Y connection. This is an important advantage when transformers are used to interconnect systems of different voltages in a cascading manner. For example, suppose there are four systems operating at 800, 440, 220, and 66 kV that need to be interconnected. Substations can be constructed using Y-Y transformer connections to interconnect any two of these voltages. The 800 kV systems can be tied with the 66 kV systems through a single 800 to 66 kV transformation or through a series of cascading transformations at 440,220 and 66 kV.
- **Required Few Turns for winding:** Due to star connection, phase voltages is (1/√3) times the line voltage. Hence less number of turns is required. Also the stress on insulation is less. This makes the connection economical for small high voltage purposes.
- **Required Less Insulation Level:** If the neutral end of a Y-connected winding is grounded, then there is an opportunity to use reduced levels of insulation at the neutral end of the winding. A winding that is connected across the phases requires full insulation throughout the winding.
- **Handle Heavy Load:** Due to star connection, phase current is same as line current. Hence windings have to carry high currents. This makes cross section of the windings high. Thus the windings are mechanically strong and windings can bear heavy loads and short circuit current.
- **Use for Three phases Four Wires System:** As neutral is available, suitable for three phases four wire

system.

- **Eliminate Distortion in Secondary Phase Voltage:** The connection of primary neutral to the neutral of generator eliminates distortion in the secondary phase voltages by giving path to triple frequency currents toward to generator.

- **Sinusoidal voltage on secondary side:** Neutral give path to flow Triple frequency current to flow Generator side thus sinusoidal voltage on primary will give sinusoidal voltage on secondary side.

- **Used as Auto Transformer:** A Y-Y transformer may be constructed as an autotransformer, with the possibility of great cost savings compared to the two-winding transformer construction.

- **Better Protective Relaying:** The protective relay settings will be protecting better on the line to ground faults when the Y-Y transformer connections with solidly grounded neutrals are applied.

Disadvantages

- **The Third harmonic issue:** The voltages in any phase of a Y-Y transformer are 1200 apart from the voltages in any other phase. However, the third-harmonic components of each phase will be in phase with each other. Nonlinearities in the transformer core always lead to generation of third harmonic. These components will add up resulting in large (can be even larger than the fundamental component) third harmonic component.

- **Overvoltage at Lighting Load:** The presence of third (and other zero-sequence) harmonics at an ungrounded neutral can cause overvoltage conditions at light load. When constructing a Y-Y transformer using single-phase transformers connected in a bank, the measured line-to-neutral voltages are not 57.7% of the system phase-to-phase voltage at no load but are about 68% and diminish very rapidly as the bank is loaded. The effective values of voltages at different frequencies combine by taking the square root of the sum of the voltages squared. With sinusoidal phase-to-phase voltage, the third-harmonic component of the phase-to-neutral voltage is about 60%.

- **Voltage drop at Unbalance Load:** There can be a large voltage drop for unbalanced phase-to-neutral loads. This is caused by the fact that phase-to-phase loads cause a voltage drop through the leakage reactance of the transformer whereas phase-to-neutral loads cause a voltage drop through the magnetizing reactance, which is 100 to 1000 times larger than the leakage reactance.
- **Overheated Transformer Tank:** Under certain circumstances, a Y-Y connected three-phase trans- can produce severe tank overheating that can quickly destroy the transformer. This usually occurs with an open phase on the primary circuit and load on the secondary.
- **Over Excitation of Core in Fault Condition:** If a phase-to-ground fault occurs on the primary circuit with the primary neutral grounded, then the phase-to-neutral voltage on the un faulted phases increases to 173% of the normal voltage. This would almost certainly result in over excitation of the core, with greatly increased magnetizing currents and core losses
- If the neutrals of the primary and secondary are both brought out, then a phase-to-ground fault on the secondary circuit causes neutral fault current to flow in the primary circuit. Ground protection re- laying in the neutral of the primary circuit may then operate for faults on the secondary circuit
- **Neutral Shifting:** If the load on the secondary side unbalanced then the performance of this connection is not satisfactory then the shifting of neutral point is possible. To prevent this, star point of the primary is required to be connected to the star point of the generator.
- **Distortion of Secondary voltage:** Even though the star or neutral point of the primary is earthed, the third harmonic present in the alternator voltage may appear on the secondary side. This causes distortion in the secondary phase voltages.
- **Over Voltage at Light Load:** The presence of third (and other zero-sequence) harmonics at an ungrounded neutral can cause overvoltage conditions at light load.

- **Difficulty in coordination of Ground Protection:** In Y-Y Transformer, a low-side ground fault causes primary ground fault current, making coordination more difficult.
- **Increase Healthy Phase Voltage under Phase to ground Fault:** If a phase-to-ground fault occurs on the primary circuit with the primary neutral grounded, then the phase-to-neutral voltage on the UN faulted phase's increases to 173% of the normal voltage. If the neutrals of the primary and secondary are both brought out, then a phase-to-ground fault on the secondary circuit causes neutral fault current to flow in the primary circuit.
- **Trip the T/C in Line-Ground Fault:** All harmonics will propagate through the transformer, zero-sequence current path is continuous through the transformer, one line-to-ground fault will trip the transformer.
- **Suitable for Core Type Transformer:** The third harmonic voltage and current is absent in such type of connection with three phase wire system or shell type of three phase units, the third harmonic phase voltage may be high. This type of connection is more suitable for core type transformers.

Application

- This Type of Transformer is rarely used due to problems with unbalanced loads.
- It is economical for small high voltage transformers as the number of turns per phase and the amount of insulation required is less.

Star/Delta connection(Yd1/Yd11)

There is a +30 Degree or -30 Degree Phase Shift between Secondary Phase Voltage to Primary Phase Voltage. The connection of Yd1 is shown in fig. 4.14 and the voltages on primary and secondary sides is also shown on the phasor diagram. The phase angle difference between the phase voltage of high voltage side and low voltage side is -30°.

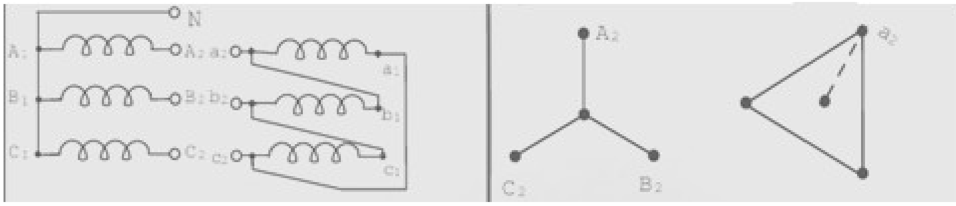


Fig 4.14. Yd1 connection and phasor diagram

The connection of Yd11 is shown in fig. 4.15 and the voltages on primary and secondary sides is also shown on the phasor diagram. The phase angle difference between the phase voltage of high voltage side and low voltage side is 30°.

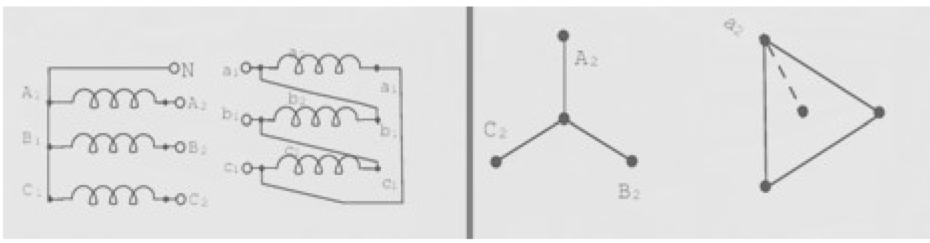


Fig 4.15. Yd11 connection and phasor diagram

Advantages

- The primary side is star connected. Hence fewer numbers of turns are required. This makes the connection economical for large high voltage step down power transformers.
- The neutral available on the primary can be earthed to avoid distortion.
- The neutral point allows both types of loads (single phase or three phases) to be met.
- Large unbalanced loads can be handled satisfactory.
- The Y-D connection has no problem with third harmonic components due to circulating currents inD. It is also more stable to unbalanced loads since the D partially redistributes any imbalance that occurs.
- The delta connected winding carries third harmonic current due to which potential of neutral point is stabilized. Some saving in cost of insulation is achieved if HV side is star connected. But in practice the HV side is normally connected in delta so that the three phase loads like motors and single phase loads like lighting loads can be supplied by LV side using three phase

four wire system.

- **As Grounding Transformer:** In Power System Mostly grounded Y- Δ transformer is used for no other purpose than to provide a good ground source in ungrounded Delta system.

Disadvantages

- In this type of connection, the secondary voltage is not in phase with the primary. Hence it is not possible to operate this connection in parallel with star-star or delta-delta connected transformer.
- One problem associated with this connection is that the secondary voltage is shifted by 30^0 with respect to the primary voltage. This can cause problems when paralleling 3-phase transformers since transformers secondary voltages must be in-phase to be paralleled. Therefore, we must pay attention to these shifts.
- If secondary of this transformer should be paralleled with secondary of another transformer without phase shift, there would be a problem

Application

- It is commonly employed for power supply transformers.
- This type of connection is commonly employed at the substation end of the transmission line. The main use with this connection is to step down the voltage. The neutral available on the primary side is grounded. It can be seen that there is phase difference of 30° between primary and secondary line voltages.
- Commonly used in a step-down transformer, Y connection on the HV side reduces insulation costs the neutral point on the HV side can be grounded, stable with respect to unbalanced loads. As for example, at the end of a transmission line. The neutral of the primary winding is earthed. In this system, line voltage ratio is $1/\sqrt{3}$ Times of transformer turn-ratio and secondary voltage lags behind primary voltage by 30°. Also third harmonic currents flows in

to give a sinusoidal flux.

Delta-star connection (Dy1/Dy11)

In this type of connection, the primary connected in delta fashion while the secondary current is connected in star. There is s +30 Degree or -30 Degree Phase Shift between Secondary Phase Voltage to Primary Phase Voltage.

The connection of Dy1 is shown in fig. 4.16 and the voltages on primary and secondary sides is also shown on the phasor diagram. The phase angle difference between the phase voltage of high voltage side and low voltage side is -30°.

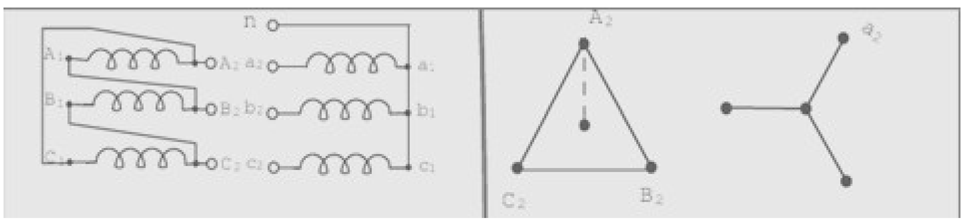


Fig 4.16. Dy1 connection and phasor diagram

The connection of Dy11 is shown in fig. 4.17 and the voltages on primary and secondary sides is also shown on the phasor diagram. The phase angle difference between the phase voltage of high voltage side and low voltage side is 30°.

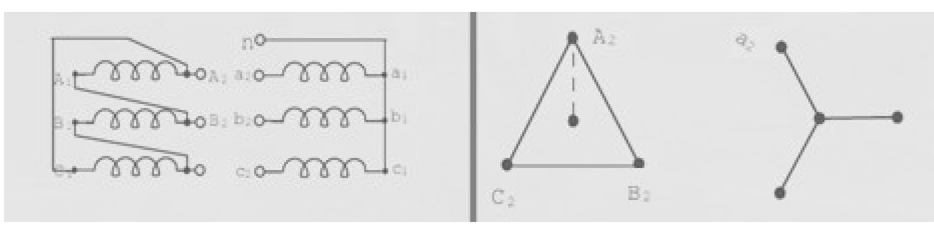


Fig 4.17. Dy11 connection and phasor diagram

Advantages

- **Cross section area of winding is less at Primary side**: On primary side due to delta connection winding cross-section required is less.

- **Used at Three phase four wire System:** On secondary side, neutral is available, due to which it can be used for 3-phase, 4 wire supply system.
- **No distortion of Secondary Voltage:** No distortion due to third harmonic components.
- **Handled large unbalanced Load:** Large unbalanced loads can be handled without any difficulty.
- **Grounding Isolation between Primary and Secondary:** Assuming that the neutral of the Y-connected secondary circuit is grounded, a load connected phase-to-neutral or a phase-to-ground fault produces two equal and opposite currents in two phases in the primary circuit without any neutral ground current in the primary circuit. Therefore, in contrast with the Y-Y connection, phase-to-ground faults or current unbalance in the secondary circuit will not affect ground protective relaying applied to the primary circuit. This feature enables proper coordination of protective devices and is a very important design consideration.
- The neutral of the Y grounded is sometimes referred to as a grounding bank, because it provides a local source of ground current at the secondary that is isolated from the primary circuit.
- **Harmonic Suppression:** The magnetizing current must contain odd harmonics for the induced voltages to be sinusoidal and the third harmonic is the dominant harmonic component. In a three-phase system the third harmonic currents of all three phases are in phase with each other because they are zero-sequence currents. In the Y-Y connection, the only path for third harmonic current is through the neutral. In the Δ -Y connection, however, the third harmonic currents, being equal in amplitude and in phase with each other, are able to circulate around the path formed by the Δ connected winding. The same thing is true for the other zero-sequence harmonics.
- **Grounding Bank:** It provides a local source of ground current at the secondary that is isolated from the primary circuit. For suppose an ungrounded generator supplies a simple radial system

through Δ-Y transformer with grounded Neutral at secondary as shown Figure. The generator can supply a single-phase-to-neutral load through the -grounded Y transformer.

Disadvantages

- In this type of connection, the secondary voltage is not in phase with the primary. Hence it is not possible to operate this connection in parallel with star-star or delta-delta connected transformer.
- One problem associated with this connection is that the secondary voltage is shifted by 30^0 with respect to the primary voltage. This can cause problems when paralleling 3-phase transformers since transformers secondary voltages must be in-phase to be paralleled. Therefore, we must pay attention to these shifts.
- If secondary of this transformer should be paralleled with secondary of another transformer without phase shift, there would be a problem.

Application

- **Commonly used in a step-up transformer:** As for example, at the beginning of a HT transmission line. In this case neutral point is stable and will not float in case of unbalanced loading. There is no distortion of flux because existence of a Δ -connection allows a path for the third-harmonic components. The line voltage ratio is √3 times of transformer turn-ratio and the secondary voltage leads the primary one by 30°. In recent years, this arrangement has become very popular for distribution system as it provides 3- Ø, 4-wire system.
- **Commonly used in commercial, industrial, and high-density residential locations**: To supply three-phase distribution systems. An example would be a distribution transformer with a delta primary, running on three 11kV phases with no neutral or earth required, and a star (or wye) secondary providing a 3-phase supply at 400 V, with the domestic voltage of 230 available between each phase and an earthed neutral point.
- **Used as Generator Transformer:** The Δ-Y transformer connection is used universally for connecting generators to transmission systems.

Delta-zigzag and Star zigzag connections (Dz0/Dz6 & Yz1/Yz6) –

The connection of Dz0 is shown in fig. 4.18 and the voltages on primary and secondary sides is also shown on the phasor diagram. The phase angle difference between the phase voltage of high voltage side and low voltage side is 0°.

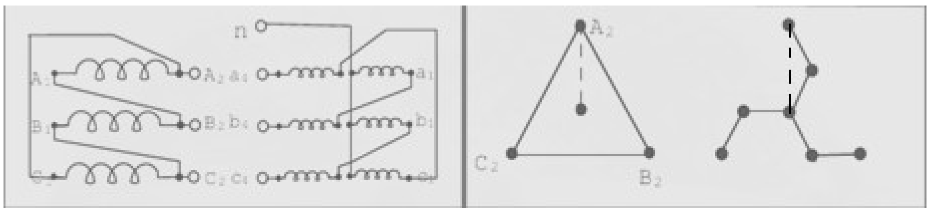


Fig 4.18. Dz0 connection and phasor diagram

The connection of Dz6 is shown in fig. 4.19 and the voltages on primary and secondary sides is also shown on the phasor diagram. The phase angle difference between the phase voltage of high voltage side and low voltage side is 180°.

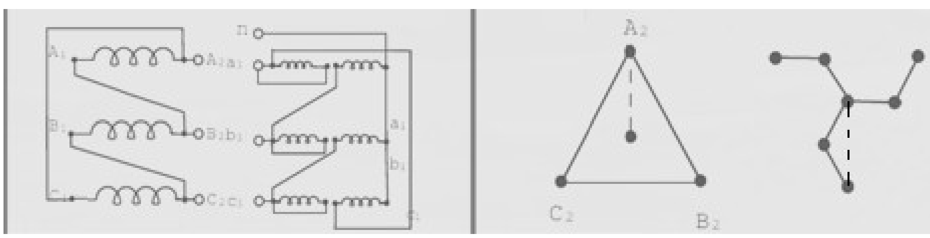


Fig 4.19. Dz6 connection and phasor diagram

The connection of Yz1 is shown in fig. 4.20 and the voltages on primary and secondary sides is also shown on the phasor diagram. The phase angle difference between the phase voltage of high voltage side and low voltage side is -30°.

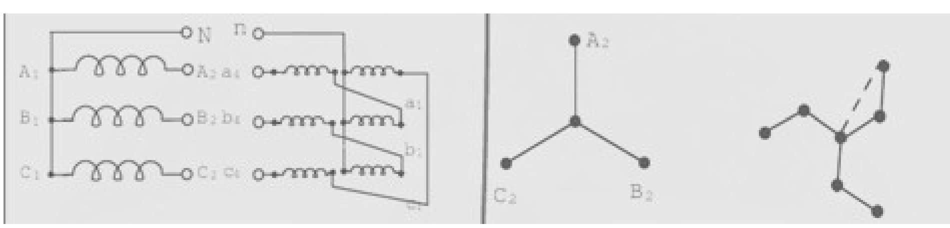


Fig 4.20. Yz1 connection and phasor diagram

The connection of Yz11 is shown in fig. 4.21 and the voltages on primary and secondary sides is also shown on the phasor diagram. The phase angle difference between the phase voltage of high voltage side and low voltage side is 30°.

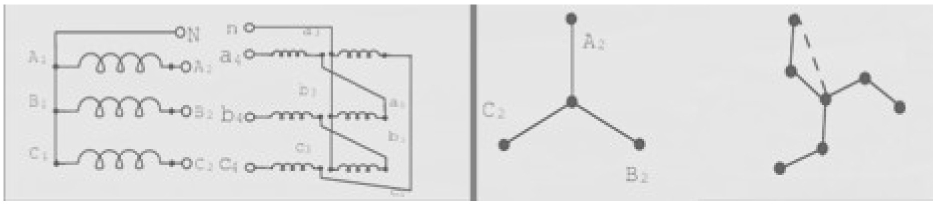


Fig 4.22 Yz11 connection and phasor diagram

- These connections are employed where delta connections are weak. Interconnection of phases in zigzag winding effects a reduction of third harmonic voltages and at the same time permits unbalanced loading.
- This connection may be used with either delta connected or star connected winding either for step-up or step-down transformers. In either case, the zigzag winding produces the same angular displacement as a delta winding, and at the same time provides a neutral for earthing purposes.
- The amount of copper required from a zigzag winding in 15% more than a corresponding star or delta winding. This is extensively used for earthing transformer.
- Due to **zigzag** connection (interconnection between phases), third harmonic voltages are reduced. It also allows unbalanced loading. The zigzag connection is employed for LV winding. For a given total voltage per phase, the zigzag side requires 15% more turns as compared to normal phase connection. In cases where delta connections are weak due to large number of turns and small cross sections, then zigzag star connection is preferred. It is also used in rectifiers.

Scott connection

There are two main reasons for the need to transform from three phases to two phases,

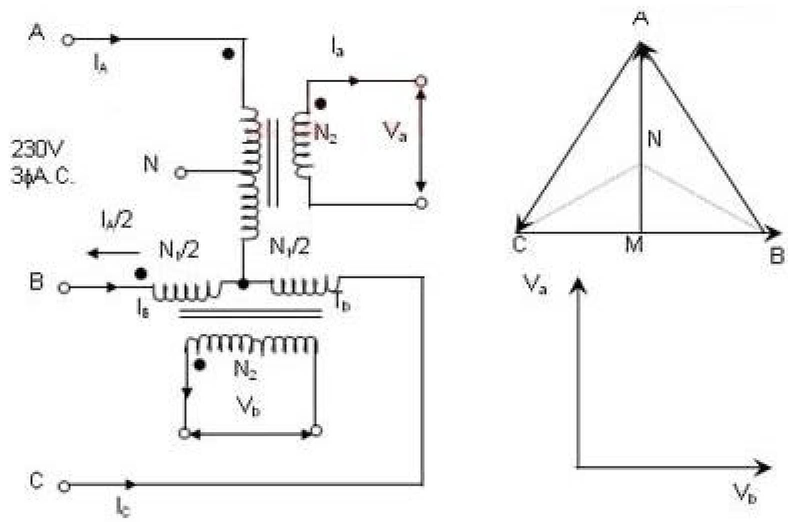
1. To give a supply to an existing two phase system from a three phase supply.

2. To supply two phase furnace transformers from a three phase source.

Two-phase systems can have 3-wire, 4-wire, or 5-wire circuits. It is needed to be considering that a two-phase system is not 2/3 of a three-phase system. Balanced three-wire, two-phase circuits have two phase wires, both carrying approximately the same amount of current, with a neutral wire carrying 1.414 times the currents in the phase wires. The phase-to-neutral voltages are 90° out of phase with each other.

Two phase 4-wire circuits are essentially just two ungrounded single-phase circuits that are electrically 90° out of phase with each other. Two phase 5-wire circuits have four phase wires plus a neutral; the four phase wires are 90° out of phase with each other.

A Scott-T transformer (also called a Scott connection) is a type of circuit used to derive two-phase power from a three-phase source or vice-versa. The Scott connection evenly distributes a balanced load between the phases of the source. Scott T Transformers require a three phase power input and provide two equal single phase outputs called Main and Teaser. The MAIN and Teaser outputs are 90 degrees out of phase. The MAIN and the Teaser outputs must not be connected in parallel or in series as it creates a vector current imbalance on the primary side. MAIN and Teaser outputs are on separate cores. An external jumper is also required to connect the primary side of the MAIN and Teaser sections. The schematic of a typical Scott T Transformer is shown below:



4.23 Connection diagram of Scott-connected transformer and vector relation of input and output

From the phasor diagram it is clear that the secondary voltages are of two phases with equal magnitude and 90° phase displacement.

Scott T Transformer is built with two single phase transformers of equal power rating. Assuming the desired voltage is the same on the two and three phase sides, the Scott-T transformer connection consists of a center-tapped 1:1 ratio main transformer, T1, and an 86.6% (0.5√3) ratio teaser transformer, T2. The center-tapped side of T1 is connected between two of the phases on the three-phase side. Its center tap then connects to one end of the lower turn count side of T2, the other end connects to the remaining phase. The other side of the transformers then connects directly to the two pairs of a two-phase four-wire system.

If the main transformer has a turn's ratio of 1: 1, then the teaser transformer requires a turn's ratio of 0.866: 1 for balanced operation. The principle of operation of the Scott connection can be most easily seen by first applying a current to the teaser secondary windings, and then applying a current to the main secondary winding, calculating the primary currents separately and superimposing the results.

The primary three-phase currents are balanced; i.e., the phase currents have the same magnitude and their phase angles are 120° apart. The apparent power supplied by the main transformer is greater than the apparent power supplied by the teaser transformer. This is easily verified by observing that the primary currents in both transformers have the same magnitude; however, the primary voltage of the teaser transformer is only 86.6% as great as the primary voltage of the main transformer. Therefore, the teaser transforms only 86.6% of the apparent power transformed by the main.

- The total real power delivered to the two phase load is equal to the total real power supplied from the three-phase system, the total apparent power transformed by both transformers is greater than the total apparent power delivered to the two-phase load.
- The apparent power transformed by the teaser is 0.866 X IH1= 1.0 and the apparent power transformed by the main is 1.0X IH2 =1.1547 for a total of 2.1547 of apparent power transformed.
- The additional 0.1547 per unit of apparent power is due to parasitic reactive power owing between the two halves of the primary winding in the main transformer.
- Single-phase transformers used in the Scott connection are specialty items that are virtually impossible to buy ‘‘off the shelf ’’ nowadays. In an emergency, standard distribution transformers can be used.

If desired, a three phase, two phase, or single phase load may be supplied simultaneously using scott-connection. The neutral points can be available for grounding or loading purposes. The Scott T connection in theory would be suitable for supplying a three, two and single phase load simultaneously, but such loads are not found together in modern practice.

The Scott T would not be recommended as a connection for 3 phase to 3 phase applications for the following reasons:

The loads of modern buildings and office buildings are inherently unbalanced and contain equipment that can be sensitive to potential voltage fluctuations that may be caused by the Scott T design.

A properly sized Scott T transformer will have to be a minimum of 7.75% larger than the equivalent Delta-Wye transformer. Properly sized, it would be a bulkier and heavier option and should not be considered a less expensive solution.

Open Delta or V-Connection

As seen previously in connection of three single phase transformers that if one of the transformers is unable to operate then the supply to the load can be continued with the remaining two transformers at the cost of reduced efficiency. The connection that obtained is called V-V connection or open delta connection.

Consider the Fig. 4.24 in which 3 phase supply is connected to the primaries. At the secondary side three equal three phase voltages will be available on no load.

The voltages are shown on phasor diagram. The connection is used when the three phase load is very very small to warrant the installation of full three phase transformer.

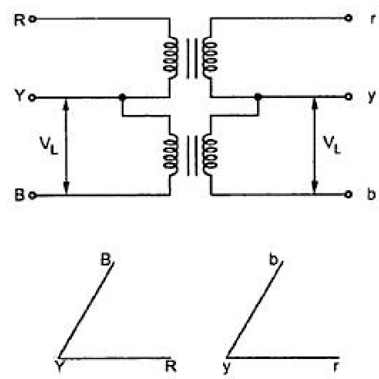


Fig. 4.24 Open delta connection of transformer at noload

If one of the transformers fails in Δ - Δ bank and if it is required to continue the supply even though at reduced capacity until the transformer which is removed from the bank is repaired or a new one is installed then this type of connection is most suitable.

When it is anticipated that in future the load increase, then it requires closing of open delta. In such cases open delta connection is preferred. It can be noted here that the removal of one of the transformers will not give the total load carried by V - V bank as two third of the capacity of Δ - Δ bank.

The load that can be carried by V - V bank is only 57.7% of it.

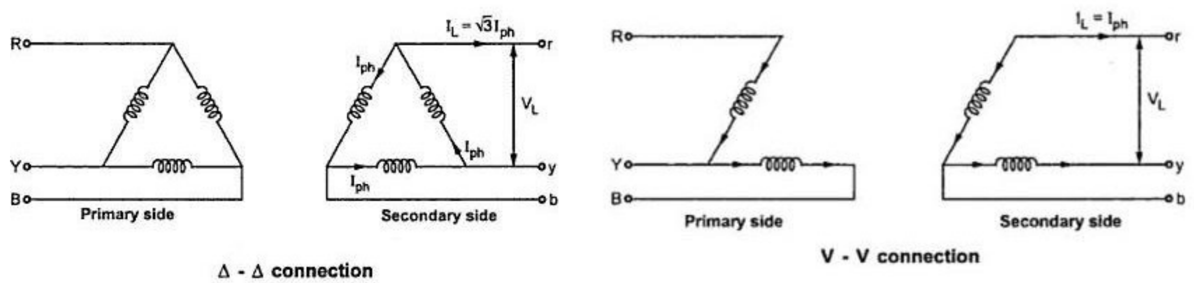


Fig. 4.25 Delta-delta and V-V connection

It can be seen from the Fig. 4.25 of delta delta connection that

Δ - Δ capacity = $\sqrt{3}\ V_L\ I_L = \sqrt{3}\ V_L\ (\sqrt{3}\ I_{ph})$

Δ - Δ capacity = $3\ V_L\ I_{ph}$

It can also be noted from the Fig. 4.25 V-V connection that the secondary line current I_L is equal to the phase current I_{ph}.

V- V capacity = $\sqrt{3}\ V_L\ I_L = \sqrt{3}\ V_L\ I_{ph}$

$$\text{So, } \frac{\text{V-V capacity}}{\Delta-\Delta \text{ capacity}} = \frac{\sqrt{3}V_L I_{ph}}{3V_L I_{ph}} = \frac{1}{\sqrt{3}} = 0.577 \simeq 58\%$$

Thus the three phase load that can be carried without exceeding the ratings of the transformers is 57.5 percent of the original load. Hence it is not 66.7 % which was expected otherwise.

The reduction in the rating can be calculated as {(66.67 - 57.735)/(57.735)}x 100 = 15.476

Suppose that we consider three transformers connected in Δ - Δ fashion and supplying their rated load. Now one transformer is removed then each of the remaining two transformers will be overloaded. The overload on each transformer will be given as,

$$\frac{\text{Total load in V-V}}{\text{VA rating of each transformer}} = \frac{\sqrt{3}V_L I_{ph}}{V_L I_{ph}} = \sqrt{3} = 1.732$$

This overload can be carried temporarily if provision is made to reduce the load otherwise overheating and breakdown of the remaining two transformers would take place.

- <u>The limitation with V -V connection are given below :</u>

The average p.f. at which V- V bank is operating is less than that with the load . This power p.f is 86.6 % of the balanced load p.f.

- The two transformers in V -V bank operate at different power factor except for balanced unity p.f .load.
- The terminals voltages available on the secondary side become unbalanced. This may happen even though load is perfectly balanced.
- Thus in summary we can say that if tow transformers are connected in V - V fashion and are loaded to rated capacity and one transformer is added to increase the total capacity by √3 or 173.2 %. Thus the increase in capacity is 73.2 % when converting from a V - V system to a Δ-Δ system.
- With a bank of tow single phase transformers connected in V-V fashion supplying a balanced 3 phase load with cosΦ asp.f., one of the transformer operate at a p.f. of cos (30-Φ) and other at cos (30+Φ). The powers of tow transformers are given by,

$$P_1 = KVA \cos(30-\Phi)$$

$$P_2 = KVA \cos(30+\Phi)$$

Oscillating Neutral

In addition to the operation of transformers on the sinusoidal supplies, the harmonic behavior becomes important as the size and rating of the transformer increases. The effects of the harmonic currents are

1. Additional copper losses due to harmonic currents
2. Increased core losses
3. Increased electro-magnetic interference with communication circuits.

On the other hand the harmonic voltages of the transformer cause

1. Increased dielectric stress on insulation
2. Electro static interference with communication circuits.

3. Resonance between winding reactance and feeder capacitance.

In the present times a greater awareness is generated by the problems of harmonic voltages and currents produced by non-linear loads like the power electronic converters. These combine with non-linear nature of transformer core and produce severe distortions in voltages and currents and increase the power loss. Thus the study of harmonics is of great practical significance in the operation of transformers.

In the case of single phase transformers connected to form three phase bank, each transformer is magnetically decoupled from the other. The flow of harmonic currents are decided by the type of the electrical connection used on the primary and secondary sides. Also, there are three fundamental voltages in the present case each displaced from the other by 120 electrical degrees. Because of the symmetry of the a.c. wave about the time axis only odd harmonics need to be considered. The harmonics which are triplen (multiples of three) behave in a similar manner as they are co-phasal or in phase in the three phases. The non-triplen harmonics behave in a similar manner to the fundamental and have ±120° phase displacement between them.

When the connection of the transformer is Yy without neutral wires both primary and secondary connected in star no closed path exists. As the triplen harmonics are always in phase, by virtue of the Y connection they get canceled in the line voltages. Non-triplen harmonics like fundamental, become 0 times phase value and appear in the line voltages. Line currents remain sinusoidal except for non-triplen harmonic currents. Flux wave in each transformer will be flat topped and the phase voltages remain peaked. The potential of the neutral is no longer steady. The star point oscillates due to the third harmonic voltages. This is termed as "oscillating neutral".

Tertiary winding

Apart from the Primary & Secondary windings, there sometimes placed a third winding in power transformers called "Tertiary Winding". Its purpose is to provide a circulating path for the harmonics (especially third harmonics) produced in the transformers along with power frequency (50Hz. third harmonic means 150 Hz oscillations). In delta-delta, delta-star and star-delta transformers all voltages are balanced and there is no floating of neutral or oscillating neutral. The floating of neutral is developed in the case star-star connection only. The transformers are sometimes constructed with three windings. The main windings are connected to form star-star connection and the third winding known as tertiary winding is used to make a closed delta connection to stabilize the neutrals of both primary and secondary circuits. The tertiary winding carries the third-harmonic currents.

Three Winding Transformers

Thus far we have looked at transformers which have one single primary winding and one single secondary winding. But the beauty of transformers is that they allow us to have more than just one winding in either the primary or secondary side. Transformers which have three winding are known commonly as **Three Winding Transformers**.

The principal of operation of a *three winding transformer* is no different from that of an ordinary transformer. Primary and secondary voltages, currents and turns ratios are all calculated the same, the difference this time is that we need to pay special attention to the voltage polarities of each coil winding, the dot convention marking the positive (or negative) polarity of the winding, when we connect them together.

Three winding transformers, also known as a three-coil, or three-winding transformer, contain one primary and two secondary coils on a common laminated core. They can be either a single-phase transformer or a three-phase transformer, (three-winding, three-phase transformer) the operation is the same.

Three Winding Transformers can also be used to provide either a step-up, a step-down, or a combination of both between the various windings. In fact a three winding transformers have two secondary windings on the same core with each one providing a different voltage or current level output.

As transformers operate on the principal of mutual induction, each individual winding of a three winding transformer supports the same number of volts per turn, therefore the volt-ampere product in each winding is the same, that is $N_P/N_S = V_P/V_S$ with any turns ratio between the individual coil windings being relative to the primary supply.

In electronic circuits, one transformer is often used to supply a variety of lower voltage levels for different components in the electronic circuitry. A typical application of three winding transformers is in power supplies and Triac Switching Converters. So a transformer have two secondary windings, each of which is electrically isolated from the others, just as it is electrically isolated from the primary. Then each of the secondary coils will produce a voltage that is proportional to its number of coil turns.

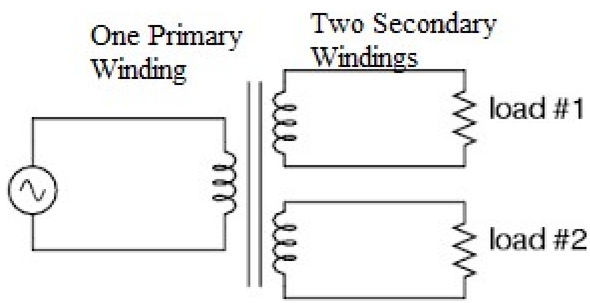


Fig. 4.27 A three winding transformer

The secondary windings can be connected together in various configurations producing a higher voltage or current supply. It must be noted that connecting together transformer windings is only possible if the two windings are electrically identical. That is their current and voltage ratings are the same.

Parallel operation of three phase transformer

4.10.1 Advantages of using transformers in parallel

1. To maximize electrical power system efficiency: Generally electrical power transformer gives the maximum efficiency at full load. If we run numbers of transformers in parallel, we can switch on only those transformers which will give the total demand by running nearer to its full load rating for that time. When load increases, we can switch none by one other transformer connected in parallel to fulfill the total demand. In this way we can run the system

with maximum efficiency.

2. To maximize electrical power system availability: If numbers of transformers run in parallel, we can shut down any one of them for maintenance purpose. Other parallel transformers in system will serve the load without total interruption of power.

3. To maximize power system reliability: If any one of the transformers run in parallel, is tripped due to fault of other parallel transformers is the system will share the load, hence power supply may not be interrupted if the shared loads do not make other transformers over loaded.

4. To maximize electrical power system flexibility: There is always a chance of increasing or decreasing future demand of power system. If it is predicted that power demand will be increased in future, there must be a provision of connecting transformers in system in parallel to fulfill the extra demand because, it is not economical from business point of view to install a bigger rated single transformer by forecasting the increased future demand as it is unnecessary investment of money. Again if future demand is decreased, transformers running in parallel can be removed from system to balance the capital investment and its return.

Conditions for parallel operation

Certain conditions have to be met before two or more transformers are connected in parallel and share a common load satisfactorily. They are,

1. The voltage ratio must be the same.
2. The per unit impedance of each machine on its own base must be the same.
3. The polarity must be the same, so that there is no circulating current between the transformers.
4. The phase sequence must be the same and no phase difference must exist between the voltages of the two transformers.

- **Same voltage ratio :** Generally the turns ratio and voltage ratio are taken to be the same. If the ratio is large there can be considerable error in the voltages even if the turns ratios are the same. When the primaries are connected to same bus bars, if the secondaries do not show the same voltage, paralleling them would result in a circulating current between the secondaries. Reflected circulating current will be there on the primary side also. Thus even without connecting a load considerable current can be drawn by the transformers and they produce copper losses. In two identical transformers with percentage impedance of 5 percent, a no-load voltage difference of one percent will result in a circulating current of 10 percent of full load current. This circulating current gets added to the load current when the load is connected resulting in unequal sharing of the load. In such cases the combined full load of the two transformers can never be met without one transformer getting overloaded.

- **Per unit impedance:** Transformers of different ratings may be required to operate in parallel. If they have to share the total load in proportion to their ratings the larger machine has to draw more current. The voltage drop across each machine has to be the same by virtue of their connection at the input and the output ends. Thus the larger machines have smaller impedance and smaller machines must have larger ohmic impedance. Thus the impedances must be in the inverse ratios of the ratings. As the voltage drops must be the same the per unit impedance of each transformer on its own base, must be equal. In addition if active and reactive power are required to be shared in proportion to the ratings the impedance angles also must be the same. Thus we have the requirement that per unit resistance and per unit reactance of both the transformers must be the same for proper load sharing.
- **Polarity of connection:** The polarity of connection in the case of single phase transformers can be either same or opposite. Inside the loop formed by the two secondaries the resulting voltage must be zero. If wrong polarity is chosen the two voltages get added and short circuit results. In the case of polyphase banks it is possible to have permanent phase error between the phases with substantial circulating current. Such transformer banks must not be connected in parallel. The turns ratios in such groups can be adjusted to give very close voltage ratios but phase errors cannot be compensated. Phase error of 0.6 degree gives rise to one percent difference in voltage. Hence poly phase transformers belonging to the same vector group alone

must be taken for paralleling.

Transformers having —30° angle can be paralleled to that having +30° angle by reversing the phase sequence of both primary and secondary terminals of one of the transformers. This way one can overcome the problem of the phase angle error.

- **Phase sequence-** The phase sequence of operation becomes relevant only in the case of poly phase systems. The poly phase banks belonging to same vector group can be connected in parallel. A transformer with +30° phase angle however can be paralleled with the one with —30° phase angle, the phase sequence is reversed for one of them both at primary and secondary terminals. If the phase sequences are not the same then the two transformers cannot be connected in parallel even if they belong to same vector group. The phase sequence can be found out by the use of a phase sequence indicator.

Load Sharing

When the transformers have equal voltage ratios, the magnitudes of secondary no-load voltages are equal. Further if the primary leakage impedance drops due to exciting currents are also equal, then $\overline{E_a} = \overline{E_b}$ and the circulating current at no load is zero.

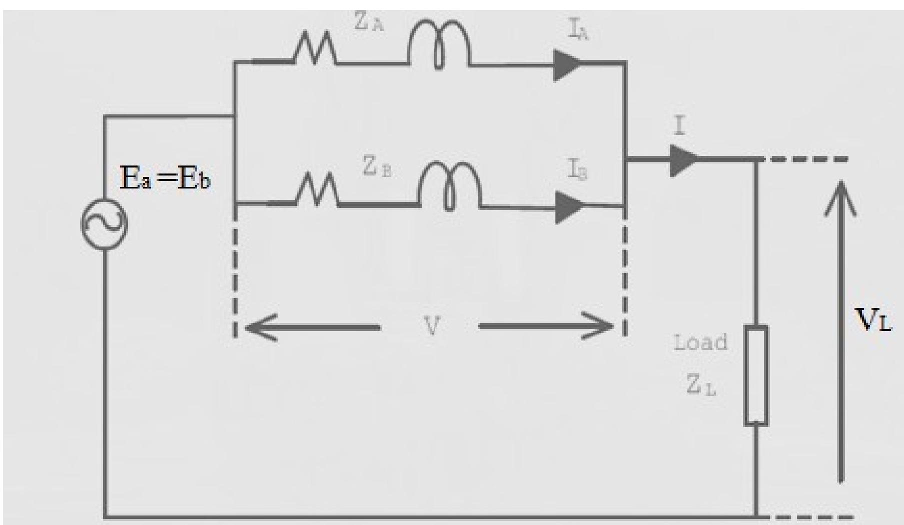


Fig. 4.28 Circuit modelling of two transformer in parallel

The equivalent circuit of two three phase transformer connected in parallel connected with a load of

Z_L impedance on per phase basis is drawn in fig 4.28. In this figure transformer A and B are operating in parallel. I_A and I_B are the load current of the two transformer.

The voltage equation of transformer A is

$$\bar{E}_a - \bar{I}_a \bar{Z}_a = \bar{V}_L = \bar{I}\bar{Z}_L$$

Since $\bar{E}_a = \bar{E}_b$; $\bar{E}_b - \bar{I}_a \bar{Z}_a = \bar{V}_L = \bar{I}\bar{Z}_L$

The voltage equation of transformer B is

$$\bar{E}_b - \bar{I}_b \bar{Z}_b = \bar{V}_L = \bar{I}\bar{Z}_L$$

$$\bar{E}_b - \bar{I}_a \bar{Z}_a = \bar{E}_b - \bar{I}_b \bar{Z}_b$$

$$\bar{I}_a \bar{Z}_a = \bar{I}_b \bar{Z}_b$$

According to the voltage drops across the two equivalent leakage impedance Za and Zb are equal.

According to KCL we can write

$$\bar{I} = \bar{I}_a + \bar{I}_b = \bar{I}_a + \frac{\bar{I}_a \bar{Z}_a}{\bar{Z}_b}$$

$$\bar{I}_a = \bar{I}\frac{\bar{Z}_b}{\bar{Z}_a + \bar{Z}_b}$$

similarly, $\bar{I}_b = \bar{I}\dfrac{\bar{Z}_a}{\bar{Z}_a + \bar{Z}_b}$

Multiplying both the current equations by terminal voltage we get,

$$\bar{S}_a = \bar{S}\frac{\bar{Z}_b}{\bar{Z}_a + \bar{Z}_b}$$

similarly, $\bar{S}_b = \bar{S}\dfrac{\bar{Z}_a}{\bar{Z}_a + \bar{Z}_b}$

Thus the power sharing in between two transformer is given in above equation in VA rating.

Acknowledgement

The committee members gratefully acknowledge google, scribd, NPTEL, openoffice, sumantra pdf, scilab for myriad suggestions and help for preparing this lecture note. The committee members also wants to express their gratitude to the persons out there who thinks knowledge should be free and be accessible and sharable without any restrictions so that every single person on this planet has the same opportunity to explore, expand and become enlightened by the collective gifts of mankind.

However apart from this lecture note students/readers are strongly recommended to follow the below mentioned books in the references and above all confer with the faculty for thorough knowledge of this authoritative subject of electrical engineering.

References

[1] P.S. Bimbhra, Electrical Machinery, Khanna Pub, 7th Ed.

[2] P.S. Bimbhra, Generalized Theory of Electrical Machine, Khanna Pub., 2nd Ed.

[3] M.G. Say, The Performance and Design of Alternating Current Machines, CBS Pub., 3rd Ed.

[4] A.E.Fitzgerald, S.D.Umans, C.Kingsley Jr., Electric Machinery, McGraw Hill Education(india) pvt. Ltd., 6th Ed.

[5] A.S. langsdorf, Theory of Alternating Current Machinery, McGraw Hill Education(india) pvt. Ltd., 2nd Ed.

[6]T.C. Lloyd, A.F. Puchstein and A.G. Conrad, Alternating-Current Machines, John Wiley& Sons pub., 3rd Ed.

[7] Irving Kosow, Electric Machinery and Transformers, Pearson India, 2nd Ed.

[8] C.L.Dawes, A course in Electrical Engineering, Vol-I, Vol-II, Mcgraw Hill Book Company Inc.

[9] P.K. Mukherjee & S. Chakravorty, Electrical Machines, Dhanpat Rai Publications, 2nd Ed.

[10] I.J Nagrath & D.P.Kothari, Electrical Machines, McGraw Hill Education(india) pvt. Ltd., 4th Ed.

[11] N.N.Parker Smith, Problems in Electrical Engineering, CBS Pub. 9th Ed.

Best of Luck to All the Students

GANDHI ACADEMY OF TECHNOLOGY AND ENGINEERING

GOLANTHARA, BERHAMPUR

LECTURE NOTES

ON

ELECTRICAL MEASUREMENTS & INSTRUMENTS

For 4th Semester

(As per Syllabus prescribed by SCTE&VT, Odisha)

Prepared By

Mr. Mahesh Kumar Mishra

(Lecturer in Electrical Engineering)

MEASURING INSTRUMENTS

1.1 Definition of instruments

An instrument is a device in which we can determine the magnitude or value of the quantity to be measured. The measuring quantity can be voltage, current, power and energy etc. Generally instruments are classified in to two categories.

Instrument

Absolute Instrument **Secondary Instrument**

1.2 Absolute instrument

An absolute instrument determines the magnitude of the quantity to be measured in terms of the instrument parameter. This instrument is really used, because each time the value of the measuring quantities varies. So we have to calculate the magnitude of the measuring quantity, analytically which is time consuming. These types of instruments are suitable for laboratory use. Example: Tangent galvanometer.

1.3 Secondary instrument

This instrument determines the value of the quantity to be measured directly. Generally these instruments are calibrated by comparing with another standard secondary instrument.

Examples of such instruments are voltmeter, ammeter and wattmeter etc. Practically secondary instruments are suitable for measurement.

Secondary instruments

Indicating instruments **Recording** **Integrating** **Electromechanically Indicating instruments**

1.3.1 Indicating instrument

This instrument uses a dial and pointer to determine the value of measuring quantity. The pointer indication gives the magnitude of measuring quantity.

1.3.2 Recording instrument

This type of instruments records the magnitude of the quantity to be measured continuously over a specified period of time.

1.3.3 Integrating instrument

This type of instrument gives the total amount of the quantity to be measured over a specified period of time.

1.3.4 Electromechanical indicating instrument

For satisfactory operation electromechanical indicating instrument, three forces are necessary.
They are

(a) Deflecting force

(b) Controlling force

(c)Damping force

1.4 Deflecting force

When there is no input signal to the instrument, the pointer will be at its zero position. To deflect the pointer from its zero position, a force is necessary which is known as deflecting force. A system which produces the deflecting force is known as a deflecting system. Generally a deflecting system converts an electrical signal to a mechanical force.

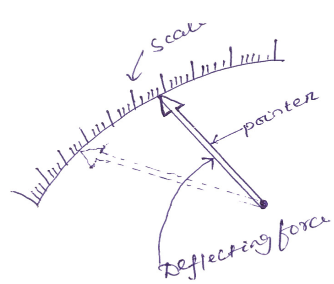


Fig. 1.1 Pointer scale

1.4.1 Magnitude effect

When a current passes through the coil (Fig.1.2), it produces a imaginary bar magnet. When a soft-iron piece is brought near this coil it is magnetized. Depending upon the current direction the poles are produced in such a way that there will be a force of attraction between the coil and the soft iron piece. This principle is used in moving iron attraction type instrument.

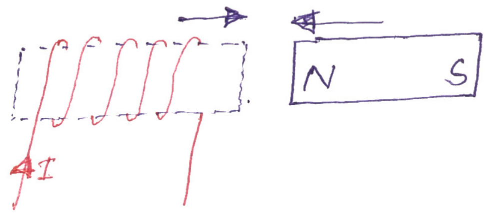


Fig. 1.2

If two soft iron pieces are place near a current carrying coil there will be a force of repulsion between the two soft iron pieces. This principle is utilized in the moving iron repulsion type instrument.

1.4.2 Force between a permanent magnet and a current carrying coil

When a current carrying coil is placed under the influence of magnetic field produced by a permanent magnet and a force is produced between them. This principle is utilized in the moving coil type instrument.

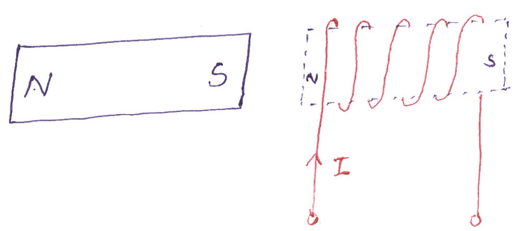


Fig. 1.3

1.4.3 Force between two current carrying coil

When two current carrying coils are placed closer to each other there will be a force of repulsion between them. If one coil is movable and other is fixed, the movable coil will move away from the fixed one. This principle is utilized in electrodynamometer type instrument.

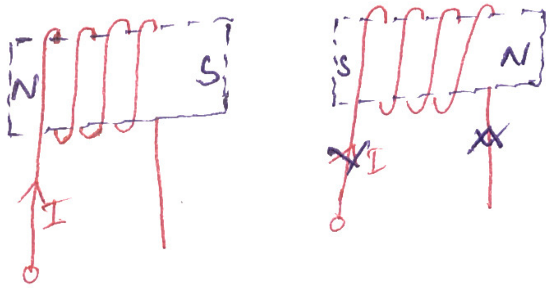


Fig. 1.4

1.5 Controlling force

To make the measurement indicated by the pointer definite (constant) a force is necessary which will be acting in the opposite direction to the deflecting force. This force is known as controlling force. A system which produces this force is known as a controlled system. When the external signal to be measured by the instrument is removed, the pointer should return back to the zero position. This is possibly due to the controlling force and the pointer will be indicating a steady value when the deflecting torque is equal to controlling torque.

$$T_d = T_c \tag{1.1}$$

1.5.1 Spring control

Two springs are attached on either end of spindle (Fig. 1.5).The spindle is placed in jewelled bearing, so that the frictional force between the pivot and spindle will be minimum. Two springs are provided in opposite direction to compensate the temperature error. The spring is made of phosphorous bronze.

When a current is supply, the pointer deflects due to rotation of the spindle. While spindle is rotate, the spring attached with the spindle will oppose the movements of the pointer. The torque produced by the spring is directly proportional to the pointer deflection θ.

$$T_C \propto \theta \tag{1.2}$$

The deflecting torque produced T_d proportional to 'I'. When $T_C = T_d$, the pointer will come to a steady position. Therefore

$$\theta \propto I \tag{1.3}$$

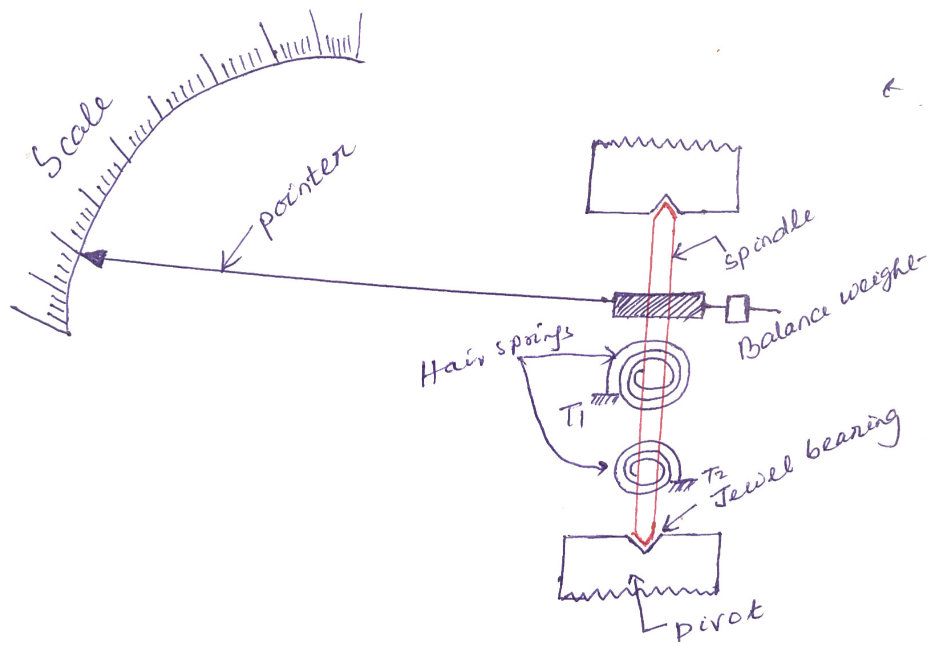


Fig. 1.5

Since, θ and I are directly proportional to the scale of such instrument which uses spring controlled is uniform.

1.6 Damping force

The deflection torque and controlling torque produced by systems are electro mechanical. Due to inertia produced by this system, the pointer oscillates about it final steady position before coming to rest. The time required to take the measurement is more. To damp out the oscillation is quickly, a damping force is necessary. This force is produced by different systems.

(a) Air friction damping
(b) Fluid friction damping
(c) Eddy current damping

1.6.1 Air friction damping

The piston is mechanically connected to a spindle through the connecting rod (Fig. 1.6). The pointer is fixed to the spindle moves over a calibrated dial. When the pointer oscillates in clockwise direction, the piston goes inside and the cylinder gets compressed. The air pushes the piston upwards and the pointer tends to move in anticlockwise direction.

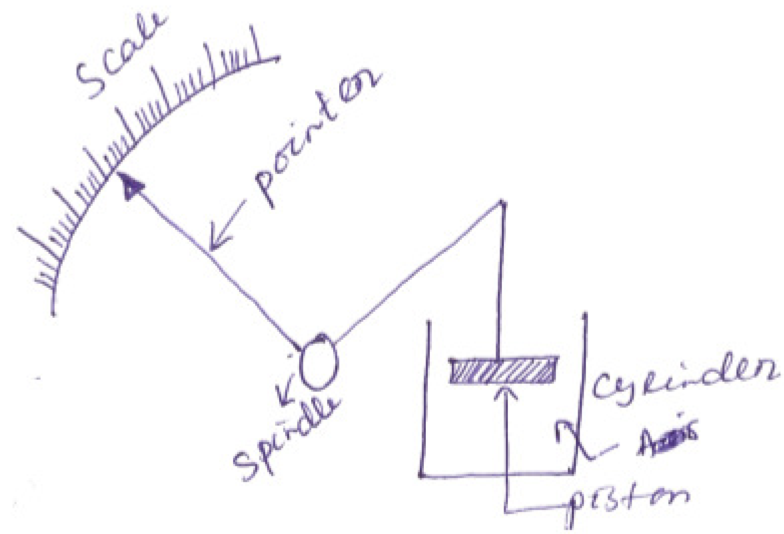


Fig. 1.6

If the pointer oscillates in anticlockwise direction the piston moves away and the pressure of the air inside cylinder gets reduced. The external pressure is more than that of the internal pressure. Therefore the piston moves down wards. The pointer tends to move in clock wise direction.

1.6.2 Eddy current damping

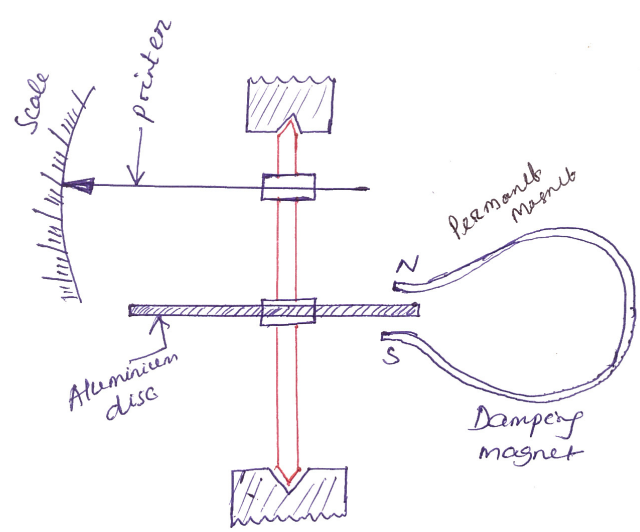


Fig. 1.6 Disc type

An aluminum circular disc is fixed to the spindle (Fig. 1.6). This disc is made to move in the magnetic field produced by a permanent magnet.

When the disc oscillates it cuts the magnetic flux produced by damping magnet. An emf is induced in the circular disc by faradays law. Eddy currents are established in the disc since it has several closed paths. By Lenz's law, the current carrying disc produced a force in a direction opposite to oscillating force. The damping force can be varied by varying the projection of the magnet over the circular disc.

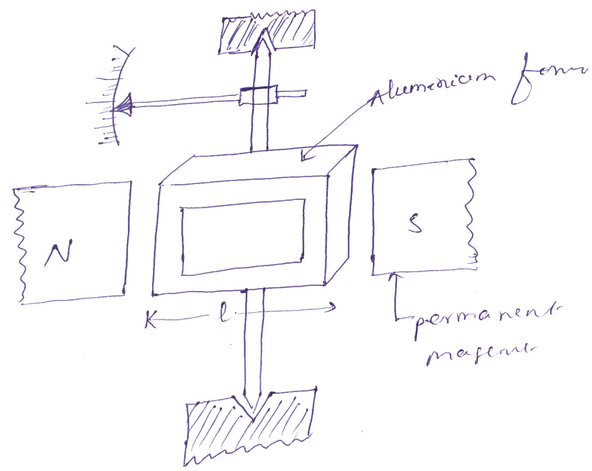


Fig. 1.6 Rectangular type

1.7 Permanent Magnet Moving Coil (PMMC) instrument

One of the most accurate type of instrument used for D.C. measurements is PMMC instrument.

Construction: A permanent magnet is used in this type instrument. Aluminum former is provided in the cylindrical in between two poles of the permanent magnet (Fig. 1.7). Coils are wound on the aluminum former which is connected with the spindle. This spindle is supported with jeweled bearing. Two springs are attached on either end of the spindle. The terminals of the moving coils are connected to the spring. Therefore the current flows through spring 1, moving coil and spring 2.

Damping: Eddy current damping is used. This is produced by aluminum former.

Control: Spring control is used.

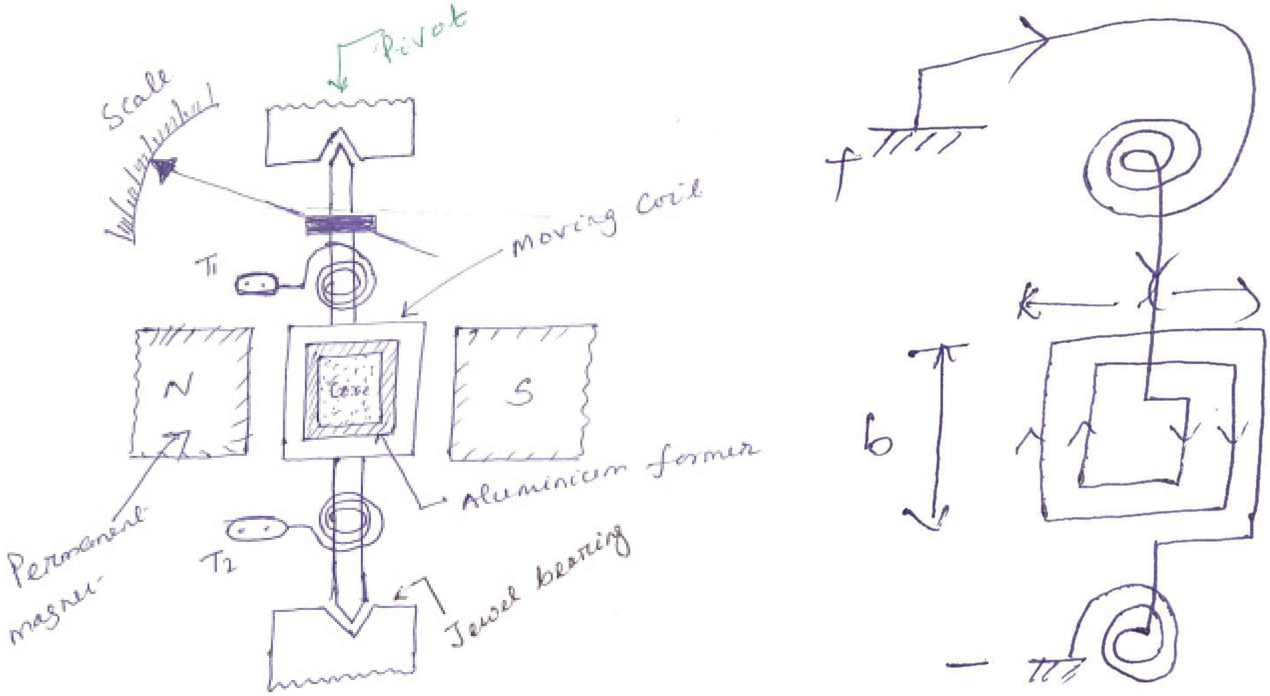


Fig. 1.7

Principle of operation

When D.C. supply is given to the moving coil, D.C. current flows through it. When the current carrying coil is kept in the magnetic field, it experiences a force. This force produces a torque and the former rotates. The pointer is attached with the spindle. When the former rotates, the pointer moves over the calibrated scale. When the polarity is reversed a torque is produced in the opposite direction. The mechanical stopper does not allow the deflection in the opposite direction. Therefore the polarity should be maintained with PMMC instrument.

If A.C. is supplied, a reversing torque is produced. This cannot produce a continuous deflection. Therefore this instrument cannot be used in A.C.

Torque developed by PMMC

Let T_d =deflecting torque

T_C = controlling torque

θ = angle of deflection

K=spring constant

b=width of the coil

l=height of the coil or length of coil

N=No. of turns

I=current

B=Flux density

A=area of the coil

The force produced in the coil is given by

$$F = BIL\sin\theta \tag{1.4}$$

When $\theta = 90^{\circ}$

For N turns, $F = NBIL$ (1.5)

Torque produced $T_d = F \times \perp_r$ distance (1.6)

$$T_d = NBIL \times b = BINA \tag{1.7}$$

$$T_d = BANI \tag{1.8}$$

$$T_d \propto I \tag{1.9}$$

Advantages

- ✓ Torque/weight is high
- ✓ Power consumption is less
- ✓ Scale is uniform
- ✓ Damping is very effective
- ✓ Since operating field is very strong, the effect of stray field is negligible
- ✓ Range of instrument can be extended

Disadvantages

- ✓ Use only for D.C.
- ✓ Cost is high
- ✓ Error is produced due to ageing effect of PMMC
- ✓ Friction and temperature error are present

1.7.1 Extension of range of PMMC instrument

Case-I: Shunt

A low shunt resistance connected in parrel with the ammeter to extent the range of current. Large current can be measured using low current rated ammeter by using a shunt.

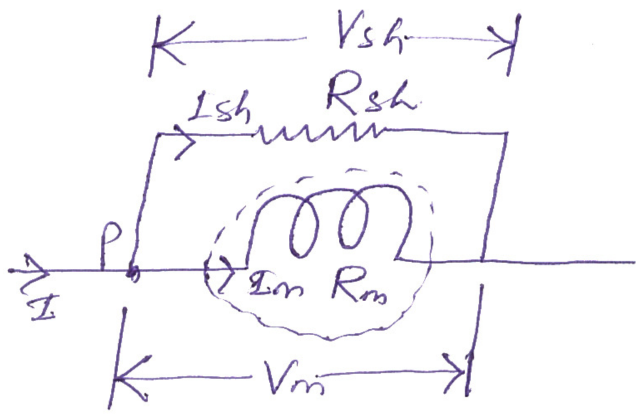


Fig. 1.8

Let R_m =Resistance of meter

R_{sh}=Resistance of shunt

I_m= Current through meter

I_{sh}=current through shunt

I= current to be measure

$\therefore V_m = V_{sh}$ (1.10)

$I_m R_m = I_{sh} R_{sh}$

$$\frac{I_m}{I_{sh}} = \frac{R_{sh}}{R_m} \quad (1.11)$$

Apply KCL at 'P' $I = I_m + I_{sh}$ (1.12)

Eq^n (1.12) ÷ by I_m

$$\frac{I}{I_m} = 1 + \frac{I_{sh}}{I_m} \quad (1.13)$$

$$\frac{I}{I_m} = 1 + \frac{R_m}{R_{sh}} \tag{1.14}$$

$$\therefore I = I_m\left(1 + \frac{R_m}{R_{sh}}\right) \tag{1.15}$$

$\left(1 + \frac{R_m}{R_{sh}}\right)$ is called multiplication factor

Shunt resistance is made of manganin. This has least thermoelectric emf. The change is resistance, due to change in temperature is negligible.

Case (II): Multiplier

A large resistance is connected in series with voltmeter is called multiplier (Fig. 1.9). A large voltage can be measured using a voltmeter of small rating with a multiplier.

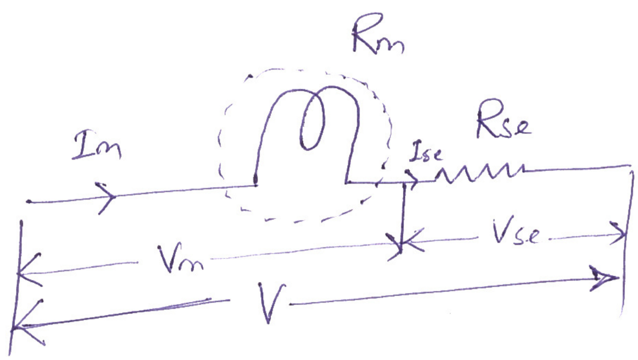


Fig. 1.9

Let R_m =resistance of meter

R_{se} =resistance of multiplier

V_m =Voltage across meter

V_{se} = Voltage across series resistance

V= voltage to be measured

$$I_m = I_{se} \tag{1.16}$$

$$\frac{V_m}{R_m} = \frac{V_{se}}{R_{se}} \tag{1.17}$$

$$\therefore \frac{V_{se}}{V_m} = \frac{R_{se}}{R_m} \tag{1.18}$$

Apply KVL, $V = V_m + V_{se}$ (1.19)

Eq^n (1.19) $\div V_m$

$$\frac{V}{V_m} = 1 + \frac{V_{se}}{V_m} = \left(1 + \frac{R_{se}}{R_m}\right) \tag{1.20}$$

$$\therefore V = V_m\left(1 + \frac{R_{se}}{R_m}\right) \tag{1.21}$$

$\left(1 + \frac{R_{se}}{R_m}\right) \rightarrow$ Multiplication factor

1.8 Moving Iron (MI) instruments

One of the most accurate instrument used for both AC and DC measurement is moving iron instrument. There are two types of moving iron instrument.

- Attraction type
- Repulsion type

1.8.1 Attraction type M.I. instrument

Construction: The moving iron fixed to the spindle is kept near the hollow fixed coil (Fig. 1.10). The pointer and balance weight are attached to the spindle, which is supported with jeweled bearing. Here air friction damping is used.

Principle of operation

The current to be measured is passed through the fixed coil. As the current is flow through the fixed coil, a magnetic field is produced. By magnetic induction the moving iron gets magnetized. The north pole of moving coil is attracted by the south pole of fixed coil. Thus the deflecting force is produced due to force of attraction. Since the moving iron is attached with the spindle, the spindle rotates and the pointer moves over the calibrated scale. But the force of attraction depends on the current flowing through the coil.

Torque developed by M.I

Let 'θ' be the deflection corresponding to a current of 'i' amp

Let the current increases by di, the corresponding deflection is '$\theta + d\theta$'

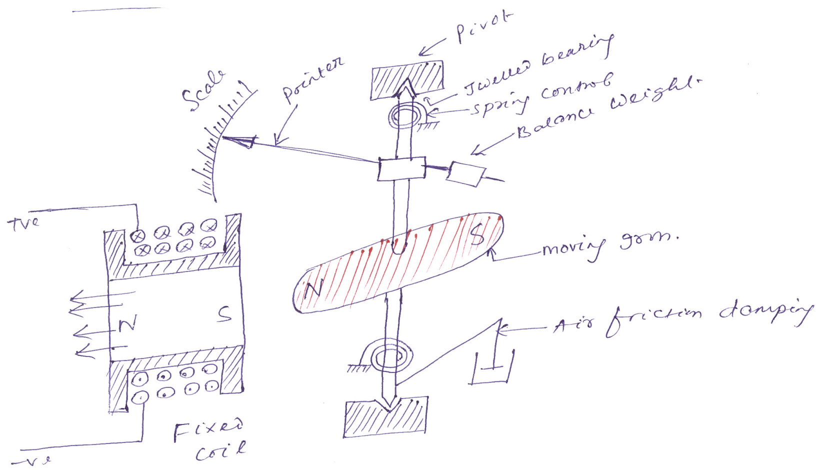


Fig. 1.10

There is change in inductance since the position of moving iron change w.r.t the fixed electromagnets.

Let the new inductance value be 'L+dL'. The current change by 'di' is dt seconds.

Let the emf induced in the coil be 'e' volt.

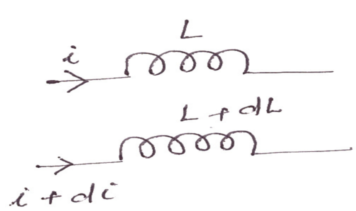
$$e=\frac{d}{dt}(Li)=L\frac{di}{dt}+i\frac{dL}{dt} \tag{1.22}$$

Multiplying by 'idt' in equation (1.22)

$$e\times idt=L\frac{di}{dt}\times idt+i\frac{dL}{dt}\times idt \tag{1.23}$$

$$e\times idt=Lidi+i^2dL \tag{1.24}$$

Eq^n (1.24) gives the energy is used in to two forms. Part of energy is stored in the inductance. Remaining energy is converted in to mechanical energy which produces deflection.



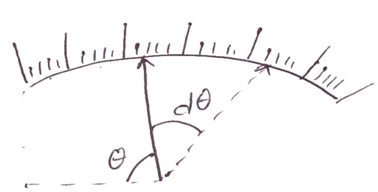


Fig. 1.11

Change in energy stored=Final energy-initial energy stored

$$=\frac{1}{2}(L+dL)(i+di)^2-\frac{1}{2}Li^2$$

$$=\frac{1}{2}\{(L+dL)(i^2+di^2+2idi)-Li^2\}$$

$$=\frac{1}{2}\{(L+dL)(i^2+2idi)-Li^2\}$$

$$=\frac{1}{2}\{Li^2+2Lidi+i^2dL+2ididL-Li^2\}$$

$$=\frac{1}{2}\{2Lidi+i^2dL\}$$

$$=Lidi+\frac{1}{2}i^2dL \qquad (1.25)$$

Mechanical work to move the pointer by $d\theta$

$$=T_d d\theta \qquad (1.26)$$

By law of conservation of energy,
Electrical energy supplied=Increase in stored energy+ mechanical work done.

Input energy= Energy stored + Mechanical energy

$$Lidi+i^2dL=Lidi+\frac{1}{2}i^2dL+T_d d\theta \qquad (1.27)$$

$$\frac{1}{2}i^2dL=T_d d\theta \qquad (1.28)$$

$$T_d=\frac{1}{2}i^2\frac{dL}{d\theta} \qquad (1.29)$$

At steady state condition $T_d=T_C$

$$\frac{1}{2}i^2\frac{dL}{d\theta}=K\theta \qquad (1.30)$$

$$\theta=\frac{1}{2K}i^2\frac{dL}{d\theta} \qquad (1.31)$$

$$\theta\propto i^2 \qquad (1.32)$$

When the instruments measure AC, $\theta\propto i^2{}_{rms}$

Scale of the instrument is non uniform.

Advantages

- ✓ MI can be used in AC and DC
- ✓ It is cheap
- ✓ Supply is given to a fixed coil, not in moving coil.
- ✓ Simple construction
- ✓ Less friction error.

Disadvantages

- ✓ It suffers from eddy current and hysteresis error
- ✓ Scale is not uniform
- ✓ It consumed more power
- ✓ Calibration is different for AC and DC operation

1.8.2 Repulsion type moving iron instrument

Construction: The repulsion type instrument has a hollow fixed iron attached to it (Fig. 1.12). The moving iron is connected to the spindle. The pointer is also attached to the spindle in supported with jeweled bearing.

Principle of operation: When the current flows through the coil, a magnetic field is produced by it. So both fixed iron and moving iron are magnetized with the same polarity, since they are kept in the same magnetic field. Similar poles of fixed and moving iron get repelled. Thus the deflecting torque is produced due to magnetic repulsion. Since moving iron is attached to spindle, the spindle will move. So that pointer moves over the calibrated scale.

Damping: Air friction damping is used to reduce the oscillation.

Control: Spring control is used.

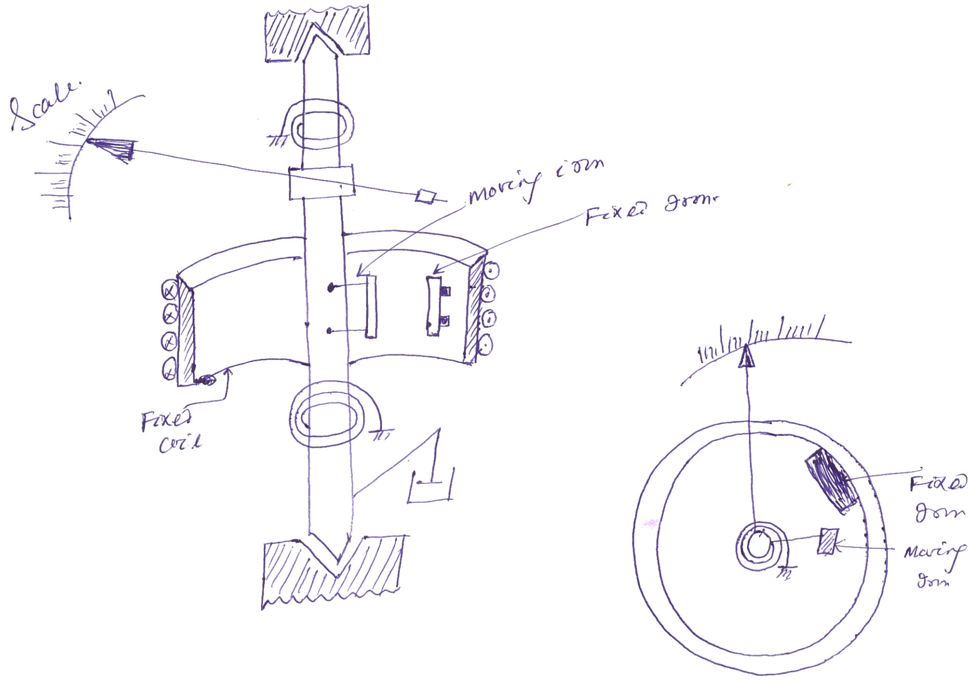


Fig. 1.12

1.9 Dynamometer (or) Electromagnetic moving coil instrument (EMMC)

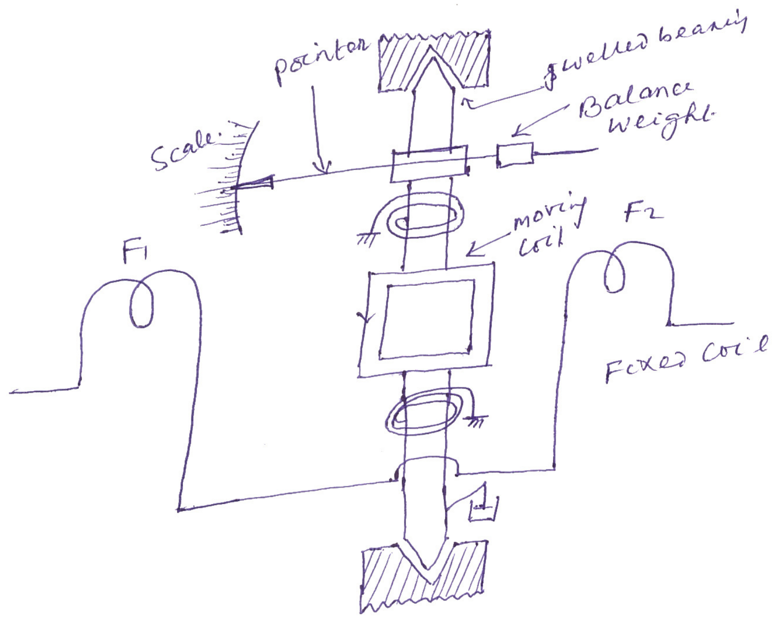


Fig. 1.13

This instrument can be used for the measurement of voltage, current and power. The difference between the PMMC and dynamometer type instrument is that the permanent magnet is replaced by an electromagnet.

Construction: A fixed coil is divided in to two equal half. The moving coil is placed between the two half of the fixed coil. Both the fixed and moving coils are air cored. So that the hysteresis effect will be zero. The pointer is attached with the spindle. In a non metallic former the moving coil is wounded.

Control: Spring control is used.

Damping: Air friction damping is used.

Principle of operation:

When the current flows through the fixed coil, it produced a magnetic field, whose flux density is proportional to the current through the fixed coil. The moving coil is kept in between the fixed coil. When the current passes through the moving coil, a magnetic field is produced by this coil.

The magnetic poles are produced in such a way that the torque produced on the moving coil deflects the pointer over the calibrated scale. This instrument works on AC and DC. When AC voltage is applied, alternating current flows through the fixed coil and moving coil. When the current in the fixed coil reverses, the current in the moving coil also reverses. Torque remains in the same direction. Since the current i_1 and i_2 reverse simultaneously. This is because the fixed and moving coils are either connected in series or parallel.

Torque developed by EMMC

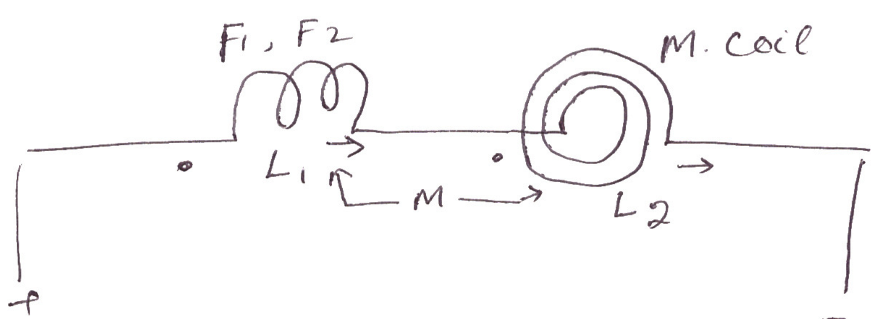


Fig. 1.14

Let

L_1=Self inductance of fixed coil

L_2= Self inductance of moving coil

M=mutual inductance between fixed coil and moving coil

i_1=current through fixed coil

i_2=current through moving coil

Total inductance of system,

$$L_{total} = L_1 + L_2 + 2M \quad (1.33)$$

But we know that in case of M.I

$$T_d = \frac{1}{2} i^2 \frac{d(L)}{d\theta} \quad (1.34)$$

$$T_d = \frac{1}{2} i^2 \frac{d}{d\theta}(L_1 + L_2 + 2M) \quad (1.35)$$

The value of L_1 and L_2 are independent of 'θ' but 'M' varies with θ

$$T_d = \frac{1}{2} i^2 \times 2 \frac{dM}{d\theta} \quad (1.36)$$

$$T_d = i^2 \frac{dM}{d\theta} \quad (1.37)$$

If the coils are not connected in series $i_1 \neq i_2$

$$\therefore T_d = i_1 i_2 \frac{dM}{d\theta} \quad (1.38)$$

$$T_C = T_d \quad (1.39)$$

$$\therefore \theta = \frac{i_1 i_2}{K} \frac{dM}{d\theta} \quad (1.40)$$

Hence the deflection of pointer is proportional to the current passing through fixed coil and moving coil.

1.9.1 Extension of EMMC instrument

Case-I Ammeter connection

Fixed coil and moving coil are connected in parallel for ammeter connection. The coils are designed such that the resistance of each branch is same.

Therefore

$I_1 = I_2 = I$

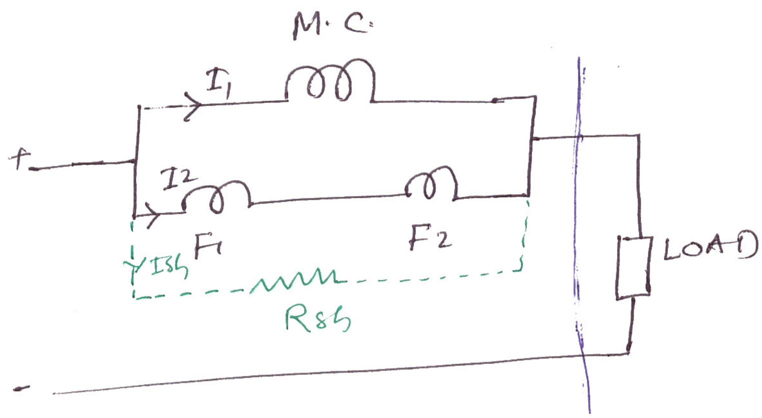


Fig. 1.15

To extend the range of current a shunt may be connected in parallel with the meter. The value R_{sh} is designed such that equal current flows through moving coil and fixed coil.

$$\therefore T_d = I_1 I_2 \frac{dM}{d\theta} \tag{1.41}$$

Or $$\therefore T_d = I^2 \frac{dM}{d\theta} \tag{1.42}$$

$$T_C = K\theta \tag{1.43}$$

$$\theta = \frac{I^2}{K}\frac{dM}{d\theta} \tag{1.44}$$

$\therefore \theta \propto I^2$ (Scale is not uniform) (1.45)

Case-II Voltmeter connection

Fixed coil and moving coil are connected in series for voltmeter connection. A multiplier may be connected in series to extent the range of voltmeter.

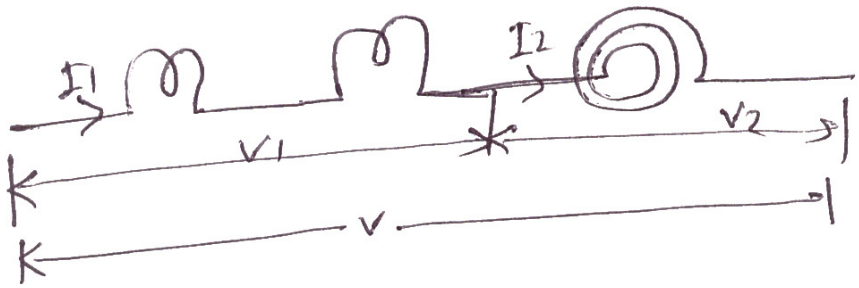


Fig. 1.16

$$I_1 = \frac{V_1}{Z_1}, I_2 = \frac{V_2}{Z_2} \tag{1.46}$$

$$T_d = \frac{V_1}{Z_1} \times \frac{V_2}{Z_2} \times \frac{dM}{d\theta} \tag{1.47}$$

$$T_d = \frac{K_1 V}{Z_1} \times \frac{K_2 V}{Z_2} \times \frac{dM}{d\theta} \tag{1.48}$$

$$T_d = \frac{KV^2}{Z_1 Z_2} \times \frac{dM}{d\theta} \tag{1.49}$$

$$T_d \propto V^2 \tag{1.50}$$

$$\therefore \theta \propto V^2 \quad \text{(Scale in not uniform)} \tag{1.51}$$

Case-III As w**attmeter**

When the two coils are connected to parallel, the instrument can be used as a wattmeter. Fixed coil is connected in series with the load. Moving coil is connected in parallel with the load. The moving coil is known as voltage coil or pressure coil and fixed coil is known as current coil.

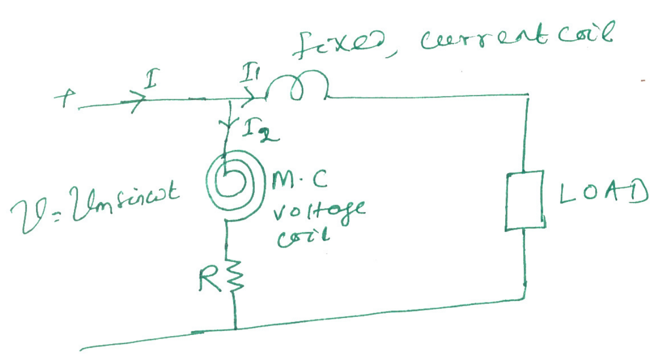


Fig. 1.17

Assume that the supply voltage is sinusoidal. If the impedance of the coil is neglected in comparison with the resistance 'R'. The current,

$$I_2 = \frac{v_m \sin wt}{R} \tag{1.52}$$

Let the phase difference between the currents I_1 and I_2 is ϕ

$$I_1 = I_m \sin(wt - \phi) \tag{1.53}$$

$$T_d = I_1 I_2 \frac{dM}{d\theta} \tag{1.54}$$

$$T_d = I_m \sin(wt - \phi) \times \frac{V_m \sin wt}{R} \frac{dM}{d\theta} \tag{1.55}$$

$$T_d = \frac{1}{R}(I_m V_m \sin wt \sin(wt - \phi)) \frac{dM}{d\theta} \tag{1.56}$$

$$T_d = \frac{1}{R} I_m V_m \sin wt . \sin(wt - \phi) \frac{dM}{d\theta} \tag{1.57}$$

The average deflecting torque

$$(T_d)_{avg} = \frac{1}{2\Pi} \int_0^{2\Pi} T_d \times d(wt) \tag{1.58}$$

$$(T_d)_{avg} = \frac{1}{2\Pi} \int_0^{2\Pi} \frac{1}{R} \times I_m V_m \sin wt . \sin(wt - \phi) \frac{dM}{d\theta} \times d(wt) \tag{1.59}$$

$$(T_d)_{avg} = \frac{V_m I_m}{2 \times 2\Pi} \times \frac{1}{R} \times \frac{dM}{d\theta} \left[\int \{ \cos\phi - \cos(2wt - \phi) \} dwt \right] \tag{1.60}$$

$$(T_d)_{avg} = \frac{V_m I_m}{4\Pi R} \times \frac{dM}{d\theta} \left[\int_0^{2\Pi} \cos\phi . dwt - \int_0^{2\Pi} \cos(2wt - \phi) . dwt \right] \tag{1.61}$$

$$(T_d)_{avg} = \frac{V_m I_m}{4\Pi R} \times \frac{dM}{d\theta} \left[\cos\phi [wt]_0^{2\Pi} \right] \tag{1.62}$$

$$(T_d)_{avg} = \frac{V_m I_m}{4\Pi R} \times \frac{dM}{d\theta} [\cos\phi (2\Pi - 0)] \tag{1.63}$$

$$(T_d)_{avg} = \frac{V_m I_m}{2} \times \frac{1}{R} \times \frac{dM}{d\theta} \times \cos\phi \tag{1.64}$$

$$(T_d)_{avg} = V_{rms} \times I_{rms} \times \cos\phi \times \frac{1}{R} \times \frac{dM}{d\theta} \tag{1.65}$$

$$(T_d)_{avg} \propto KVI\cos\phi \qquad (1.66)$$

$$T_C \propto \theta \qquad (1.67)$$

$$\theta \propto KVI\cos\phi \qquad (1.68)$$

$$\theta \propto VI\cos\phi \qquad (1.69)$$

Advantages

- ✓ It can be used for voltmeter, ammeter and wattmeter
- ✓ Hysteresis error is nill
- ✓ Eddy current error is nill
- ✓ Damping is effective
- ✓ It can be measure correctively and accurately the rms value of the voltage

Disadvantages

- ✓ Scale is not uniform
- ✓ Power consumption is high(because of high resistance)
- ✓ Cost is more
- ✓ Error is produced due to frequency, temperature and stray field.
- ✓ Torque/weight is low.(Because field strength is very low)

Errors in PMMC

- ✓ The permanent magnet produced error due to ageing effect. By heat treatment, this error can be eliminated.
- ✓ The spring produces error due to ageing effect. By heat treating the spring the error can be eliminated.
- ✓ When the temperature changes, the resistance of the coil vary and the spring also produces error in deflection. This error can be minimized by using a spring whose temperature co-efficient is very low.

1.10 Difference between attraction and repulsion type instrument

An attraction type instrument will usually have a lower inductance, compare to repulsion type instrument. But in other hand, repulsion type instruments are more suitable for economical production in manufacture and nearly uniform scale is more easily obtained. They are therefore much more common than attraction type.

1.11 Characteristics of meter

1.11.1 Full scale deflection current(I_{FSD})

The current required to bring the pointer to full-scale or extreme right side of the instrument is called full scale deflection current. It must be as small as possible. Typical value is between 2 μ A to 30mA.

1.11.2 Resistance of the coil(R_m)

This is ohmic resistance of the moving coil. It is due to ρ, L and A. For an ammeter this should be as small as possible.

1.11.3 Sensitivity of the meter(S)

$$S = \frac{1}{I_{FSD}}(\Omega / volt), \uparrow S = \frac{Z \uparrow}{V}$$

It is also called ohms/volt rating of the instrument. Larger the sensitivity of an instrument, more accurate is the instrument. It is measured in Ω/volt. When the sensitivity is high, the impedance of meter is high. Hence it draws less current and loading affect is negligible. It is also defend as one over full scale deflection current.

1.12 Error in M.I instrument

1.12.1 Temperature error

Due to temperature variation, the resistance of the coil varies. This affects the deflection of the instrument. The coil should be made of manganin, so that the resistance is almost constant.

1.12.2 Hysteresis error

Due to hysteresis affect the reading of the instrument will not be correct. When the current is decreasing, the flux produced will not decrease suddenly. Due to this the meter reads a higher value of current. Similarly when the current increases the meter reads a lower value of current. This produces error in deflection. This error can be eliminated using small iron parts with narrow hysteresis loop so that the demagnetization takes place very quickly.

1.12.3 Eddy current error

The eddy currents induced in the moving iron affect the deflection. This error can be reduced by increasing the resistance of the iron.

1.12.4 Stray field error

Since the operating field is weak, the effect of stray field is more. Due to this, error is produced in deflection. This can be eliminated by shielding the parts of the instrument.

1.12.5 Frequency error

When the frequency changes the reactance of the coil changes.

$$Z = \sqrt{(R_m + R_S)^2 + X_L^2} \tag{1.70}$$

$$I = \frac{V}{Z} = \frac{V}{\sqrt{(R_m + R_S)^2 + X_L^2}} \tag{1.71}$$

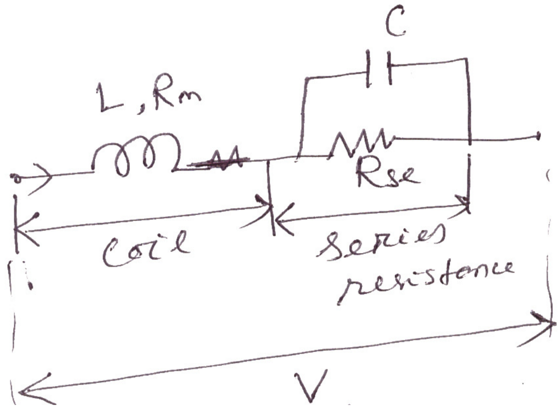


Fig. 1.18

Deflection of moving iron voltmeter depends upon the current through the coil. Therefore, deflection for a given voltage will be less at higher frequency than at low frequency. A capacitor is connected in parallel with multiplier resistance. The net reactance, ($X_L - X_C$) is very small, when compared to the series resistance. Thus the circuit impedance is made independent of frequency. This is because of the circuit is almost resistive.

$$C = 0.41 \frac{L}{(R_S)^2} \tag{1.72}$$

1.13 Electrostatic instrument

In multi cellular construction several vans and quadrants are provided. The voltage is to be measured is applied between the vanes and quadrant. The force of attraction between the vanes and quadrant produces a deflecting torque. Controlling torque is produced by spring control. Air friction damping is used.

The instrument is generally used for measuring medium and high voltage. The voltage is reduced to low value by using capacitor potential divider. The force of attraction is proportional to the square of the voltage.

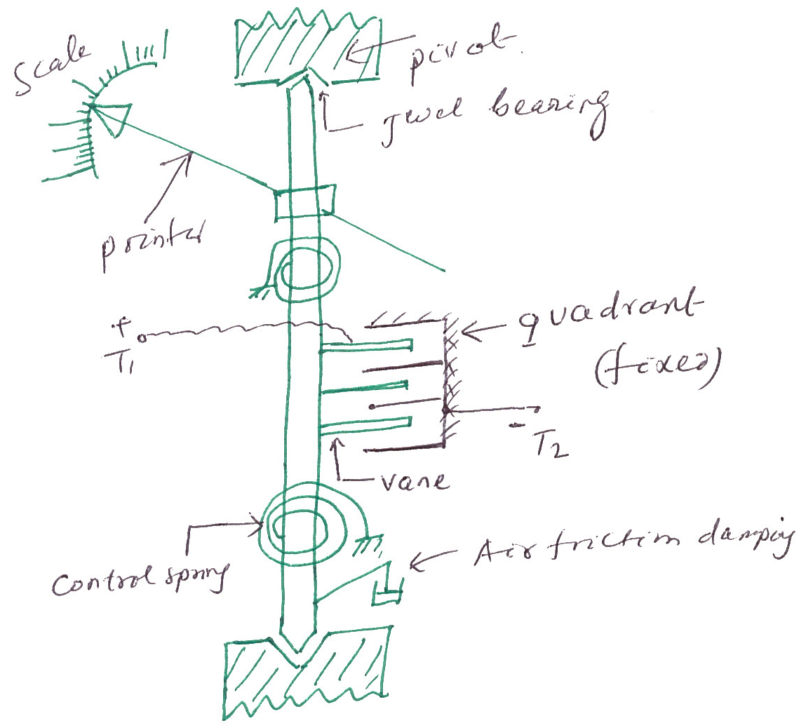


Fig. 1.19

<u>Torque develop by electrostatic instrument</u>

V=Voltage applied between vane and quadrant

C=capacitance between vane and quadrant

Energy stored= $\frac{1}{2}CV^2$ (1.73)

Let 'θ' be the deflection corresponding to a voltage V.

Let the voltage increases by dv, the corresponding deflection is' $\theta + d\theta$'

When the voltage is being increased, a capacitive current flows

$$i = \frac{dq}{dt} = \frac{d(CV)}{dt} = \frac{dC}{dt}V + C\frac{dV}{dt} \tag{1.74}$$

$V \times dt$ multiply on both side of equation (1.74)

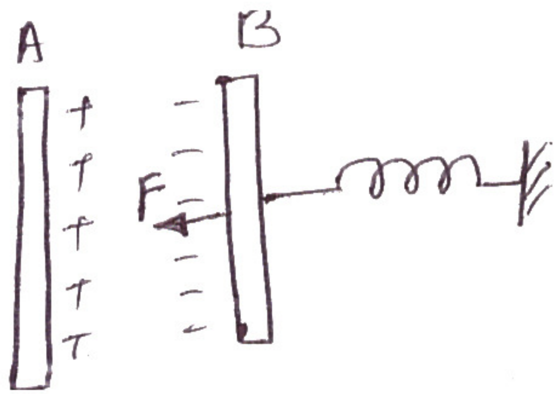


Fig. 1.20

$$Vidt = \frac{dC}{dt}V^2 dt + CV\frac{dV}{dt}dt \tag{1.75}$$

$$Vidt = V^2 dC + CVdV \tag{1.76}$$

$$\text{Change in stored energy}=\frac{1}{2}(C+dC)(V+dV)^2 - \frac{1}{2}CV^2 \tag{1.77}$$

$$= \frac{1}{2}\left[(C+dC)V^2 + dV^2 + 2VdV\right] - \frac{1}{2}CV^2$$

$$= \frac{1}{2}\left[CV^2 + CdV^2 + 2CVdV + V^2dC + dCdV^2 + 2VdVdC\right] - \frac{1}{2}CV^2$$

$$= \frac{1}{2}V^2dC + CVdV$$

$$V^2dC + CVdV = \frac{1}{2}V^2dC + CVdV + F \times rd\theta \tag{1.78}$$

$$T_d \times d\theta = \frac{1}{2}V^2dC \tag{1.79}$$

$$T_d = \frac{1}{2}V^2\left(\frac{dC}{d\theta}\right) \tag{1.80}$$

At steady state condition, $T_d = T_C$

$$K\theta = \frac{1}{2}V^2\left(\frac{dC}{d\theta}\right) \tag{1.81}$$

$$\theta = \frac{1}{2K}V^2\left(\frac{dC}{d\theta}\right) \tag{1.82}$$

Advantages

- ✓ It is used in both AC and DC.
- ✓ There is no frequency error.
- ✓ There is no hysteresis error.
- ✓ There is no stray magnetic field error. Because the instrument works on electrostatic principle.
- ✓ It is used for high voltage
- ✓ Power consumption is negligible.

Disadvantages

- ✓ Scale is not uniform
- ✓ Large in size
- ✓ Cost is more

1.14 Multi range Ammeter

When the switch is connected to position (1), the supplied current I_1

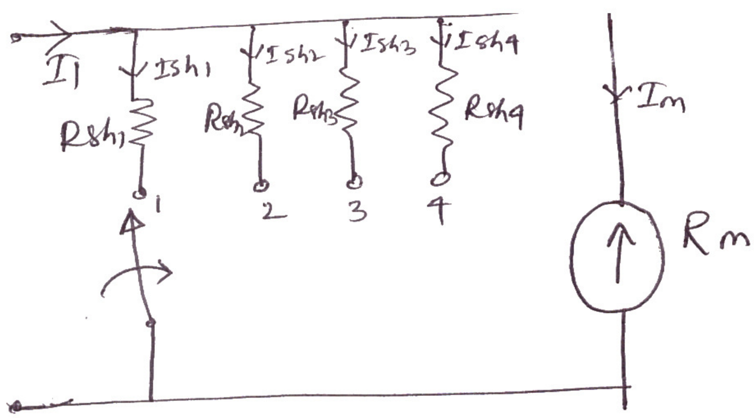


Fig. 1.21

$$I_{sh1}R_{sh1} = I_m R_m \tag{1.83}$$

$$R_{sh1} = \frac{I_m R_m}{I_{sh1}} = \frac{I_m R_m}{I_1 - I_m} \tag{1.84}$$

$$R_{sh1} = \frac{R_m}{\frac{I_1}{I_m} - 1}, R_{sh1} = \frac{R_m}{m_1 - 1}, m_1 = \frac{I_1}{I_m} = \text{ Multiplying power of shunt}$$

$$R_{sh2} = \frac{R_m}{m_2 - 1}, m_2 = \frac{I_2}{I_m} \tag{1.85}$$

$$R_{sh3} = \frac{R_m}{m_3 - 1}, m_3 = \frac{I_3}{I_m} \tag{1.86}$$

$$R_{sh4} = \frac{R_m}{m_4 - 1}, m_4 = \frac{I_4}{I_m} \tag{1.87}$$

1.15 Ayrton shunt

$$R_1 = R_{sh1} - R_{sh2} \tag{1.88}$$

$$R_2 = R_{sh2} - R_{sh3} \tag{1.89}$$

$$R_3 = R_{sh3} - R_{sh4} \tag{1.90}$$

$$R_4 = R_{sh4} \tag{1.91}$$

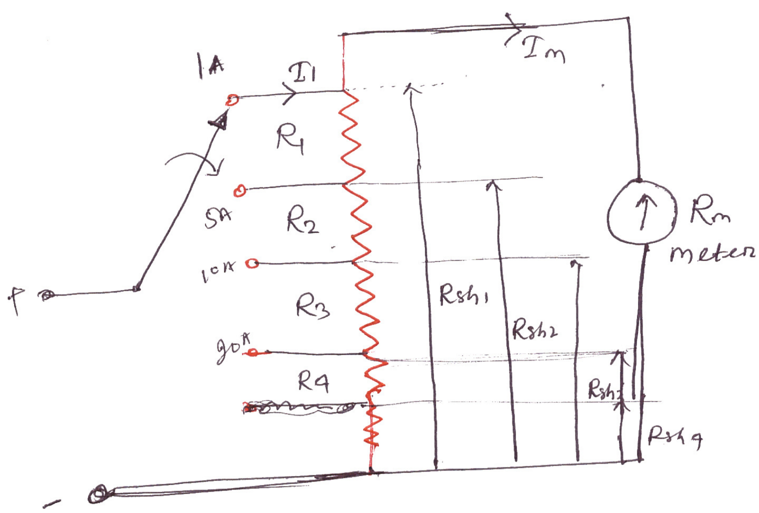


Fig. 1.22

Ayrton shunt is also called universal shunt. Ayrton shunt has more sections of resistance. Taps are brought out from various points of the resistor. The variable points in the o/p can be connected to any position. Various meters require different types of shunts. The Aryton shunt is used in the lab, so that any value of resistance between minimum and maximum specified can be used. It eliminates the possibility of having the meter in the circuit without a shunt.

1.16 Multi range D.C. voltmeter

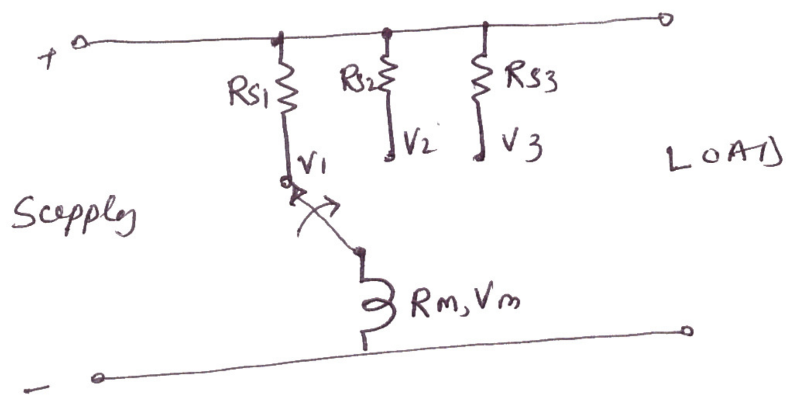


Fig. 1.23

$$R_{s1} = R_m(m_1 - 1)$$
$$R_{s2} = R_m(m_2 - 1) \quad (1.92)$$
$$R_{s3} = R_m(m_3 - 1)$$

$$m_1 = \frac{V_1}{V_m}, m_2 = \frac{V_2}{V_m}, m_3 = \frac{V_3}{V_m} \quad (1.93)$$

We can obtain different Voltage ranges by connecting different value of multiplier resistor in series with the meter. The number of these resistors is equal to the number of ranges required**.**

1.17 Potential divider arrangement

The resistance R_1, R_2, R_3 and R_4 is connected in series to obtained the ranges V_1, V_2, V_3 and V_4

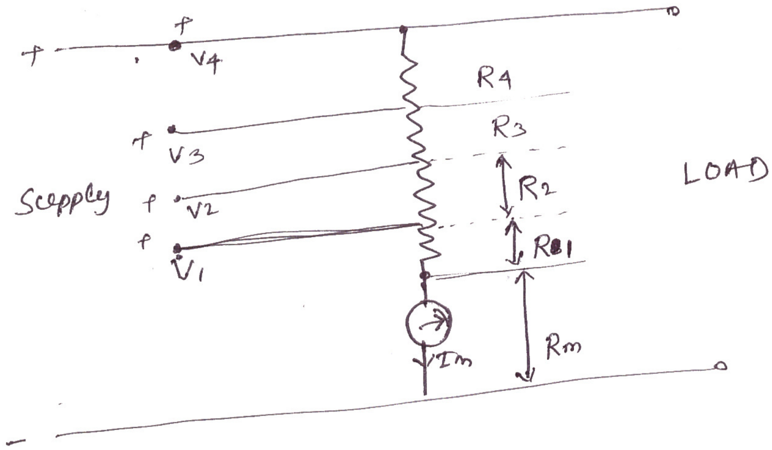


Fig. 1.24

Consider for voltage V_1, $(R_1 + R_m)I_m = V_1$

$$\therefore R_1 = \frac{V_1}{I_m} - R_m = \frac{V_1}{(\frac{V_m}{R_m})} - R_m = \left(\frac{V_1}{V_m}\right)R_m - R_m \tag{1.94}$$

$$R_1 = (m_1 - 1)R_m \tag{1.95}$$

For V_2, $(R_2 + R_1 + R_m)I_m = V_2 \Rightarrow R_2 = \frac{V_2}{I_m} - R_1 - R_m$ (1.96)

$$R_2 = \frac{V_2}{\left(\frac{V_m}{R_m}\right)} - (m_1 - 1)R_m - R_m \tag{1.97}$$

$$R_2 = m_2 R_m - R_m - (m_1 - 1)R_m$$

$$= R_m(m_2 - 1 - m_1 + 1) \tag{1.98}$$

$$R_2 = (m_2 - m_1)R_m \tag{1.99}$$

For V_3 $(R_3 + R_2 + R_1 + R_m)I_m = V_3$

$$R_3 = \frac{V_3}{I_m} - R_2 - R_1 - R_m$$

$$= \frac{V_3}{V_m}R_m - (m_2 - m_1)R_m - (m_1 - 1)R_m - R_m$$

$$= m_3 R_m - (m_2 - m_1)R_m - (m_1 - 1)R_m - R_m$$

$$R_3 = (m_3 - m_2)R_m$$

For V_4 $(R_4 + R_3 + R_2 + R_1 + R_m)I_m = V_4$

$$R_4 = \frac{V_4}{I_m} - R_3 - R_2 - R_1 - R_m$$

$$= \left(\frac{V_4}{V_m}\right)R_m - (m_3 - m_2)R_m - (m_2 - m_1)R_m - (m_1 - 1)R_m - R_m$$

$$R_4 = R_m[m_4 - m_3 + m_2 - m_2 + m_1 - m_1 + 1 - 1]$$
$$R_4 = (m_4 - m_3)R_m$$

Example: 1.1

A PMMC ammeter has the following specification

Coil dimension are 1cm$\times$1cm. Spring constant is $0.15\times10^{-6}\,N-m/rad$, Flux density is $1.5\times10^{-3}\,wb/m^2$.Determine the no. of turns required to produce a deflection of 90^0 when a current 2mA flows through the coil.

Solution:

At steady state **condition** $T_d = T_C$

$BANI = K\theta$

$$\Rightarrow N = \frac{K\theta}{BAI}$$

A=$1\times10^{-4}m^2$

K=$0.15\times10^{-6}\,\frac{N-m}{rad}$

B=$1.5\times10^{-3}wb/m^2$

I=$2\times10^{-3}A$

$\theta = 90^{\circ} = \frac{\Pi}{2}rad$

N=785 ans.

Example: 1.2

The pointer of a moving coil instrument gives full scale deflection of 20mA. The potential difference across the meter when carrying 20mA is 400mV.The instrument to be used is 200A for full scale deflection. Find the shunt resistance required to achieve this, if the instrument to be used as a voltmeter for full scale reading with 1000V. Find the series resistance to be connected it?

Solution:

Case-1

V_m=400mV

$I_m = 20mA$

I=200A

$$R_m = \frac{V_m}{I_m} = \frac{400}{20} = 20\Omega$$

$$I = I_m\left(1 + \frac{R_m}{R_{sh}}\right)$$

$$200 = 20\times10^{-3}\left[1 + \frac{20}{R_{sh}}\right]$$

$$R_{sh} = 2\times10^{-3}\Omega$$

Case-II

V=1000V

$$V = V_m\left(1 + \frac{R_{se}}{R_m}\right)$$

$$4000 = 400\times10^{-3}\left(1 + \frac{R_{se}}{20}\right)$$

$$R_{se} = 49.98k\Omega$$

Example: 1.3

A 150 v moving iron voltmeter is intended for 50HZ, has a resistance of 3kΩ. Find the series resistance required to extent the range of instrument to 300v. If the 300V instrument is used to measure a d.c. voltage of 200V. Find the voltage across the meter?

Solution:

$R_m = 3k\Omega, V_m = 150V, V = 300V$

$$V = V_m\left(1+\frac{R_{se}}{R_m}\right)$$

$$300 = 150\left(1+\frac{R_{se}}{3}\right) \Rightarrow R_{se} = 3k\Omega$$

Case-II $$V = V_m\left(1+\frac{R_{se}}{R_m}\right)$$

$$200 = V_m\left(1+\frac{3}{3}\right)$$

$\therefore V_m = 100V$ Ans

Example: 1.4

What is the value of series resistance to be used to extent '0'to 200V range of 20,000Ω/volt voltmeter to 0 to 2000 volt?

Solution:

$$V_{se} = V - V = 1800$$

$$I_{FSD} = \frac{1}{20000} = \frac{1}{Sensitivity}$$

$V_{se} = R_{se} \times i_{FSD} \Rightarrow R_{se} = 36M\Omega$ ans.

Example: 1.5

A moving coil instrument whose resistance is 25Ω gives a full scale deflection with a current of 1mA. This instrument is to be used with a manganin shunt, to extent its range to 100mA. Calculate the error caused by a 10^0C rise in temperature when:

(a) Copper moving coil is connected directly across the manganin shunt.

(b) A 75 ohm manganin resistance is used in series with the instrument moving coil.

The temperature co-efficient of copper is $0.004/^0$C and that of manganin is 0.00015^0/C.

Solution:

Case-1

$$I_m = 1mA$$

$$R_m = 25\Omega$$

I=100mA

$$I = I_m\left(1+\frac{R_m}{R_{sh}}\right)$$

$$100 = 1\left(1+\frac{25}{R_{sh}}\right) \Rightarrow \frac{25}{R_{sh}} = 99$$

$$\Rightarrow R_{sh} = \frac{25}{99} = 0.2525\Omega$$

Instrument resistance for 10^0C rise in temperature, $R_{mt} = 25(1+0.004\times 10)$

$$R_t = R_o(1+\rho_t \times t)$$

$$R_{m/t=10^\circ} = 26\Omega$$

Shunt resistance for 10^0C, rise in temperature

$$R_{sh/t=10^\circ} = 0.2525(1+0.00015\times 10) = 0.2529\Omega$$

Current through the meter for 100mA in the main circuit for 10^0C rise in temperature

$$I = I_m\left(1+\frac{R_m}{R_{sh}}\right)\Bigg|_{t=10^\circ C}$$

$$100 = I_{mt}\left(1+\frac{26}{0.2529}\right)$$

$$I_m\big|_{t=10} = 0.963mA$$

But normal meter current=1mA

Error due to rise in temperature=(0.963-1)*100=-3.7%

Case-b As voltmeter

Total resistance in the meter circuit= $R_m + R_{sh} = 25 + 75 = 100\Omega$

$$I = I_m\left(1+\frac{R_m}{R_{sh}}\right)$$

$$100 = 1\left(1+\frac{100}{R_{sh}}\right)$$

$$R_{sh} = \frac{100}{100-1} = 1.01\Omega$$

Resistance of the instrument circuit for 10^0C rise in temperature

$$R_m\big|_{t=10} = 25(1+0.004\times10)+75(1+0.00015\times10)=101.11\Omega$$

Shunt resistance for 10^0C rise in temperature

$$R_{sh}\big|_{t=10} = 1.01(1+0.00015\times10)=1.0115\Omega$$

$$I = I_m\left(1+\frac{R_m}{R_{sh}}\right)$$

$$100 = I_m\left(1+\frac{101.11}{1.0115}\right)$$

$$I_m\Big|t=10^\circ = 0.9905mA$$

Error =(0.9905-1)*100=-0.95%

Example: 1.6

The coil of a 600V M.I meter has an inductance of 1 henery. It gives correct reading at 50HZ and requires 100mA. For its full scale deflection, what is % error in the meter when connected to 200V D.C. by comparing with 200V A.C?

Solution:

$$V_m = 600V, I_m = 100mA$$

Case-I A.C.

$$Z_m = \frac{V_m}{I_m} = \frac{600}{0.1} = 6000\Omega$$

$$X_L = 2\Pi fL = 314\Omega$$

$$R_m = \sqrt{Z_m{}^2 - X_L^2} = \sqrt{(6000)^2-(314)^2} = 5990\Omega$$

$$I_{AC} = \frac{V_{AC}}{Z} = \frac{200}{6000} = 33.33mA$$

Case-II D.C

$$I_{DC} = \frac{V_{DC}}{R_m} = \frac{200}{5990} = 33.39mA$$

Error=$\frac{I_{DC}-I_{AC}}{I_{AC}}\times 100=\frac{33.39-33.33}{33.33}\times 100=0.18\%$

Example: 1.7

A 250V M.I. voltmeter has coil resistance of 500Ω, coil inductance 0f 1.04 H and series resistance of 2kΩ. The meter reads correctively at 250V D.C. What will be the value of capacitance to be used for shunting the series resistance to make the meter read correctly at 50HZ? What is the reading of voltmeter on A.C. without capacitance?

Solution: $C=0.41\frac{L}{(R_S)^2}$

$$=0.41\times\frac{1.04}{(2\times 10^3)^2}=0.1\mu F$$

For A.C $Z=\sqrt{(R_m+R_{Se})^2+X_L^2}$

$$Z=\sqrt{(500+2000)^2+(314)^2}=2520\Omega$$

With D.C

$R_{total}=2500\Omega$

For 2500Ω $\rightarrow$ 250V

$$1\Omega\rightarrow\frac{250}{2500}$$

$$2520\Omega\rightarrow\frac{250}{2500}\times 2520=248V$$

Example: 1.8

The relationship between inductance of moving iron ammeter, the current and the position of pointer is as follows:

Reading (A)	1.2	1.4	1.6	1.8
Deflection (degree)	36.5	49.5	61.5	74.5
Inductance (μH)	575.2	576.5	577.8	578.8

Calculate the deflecting torque and the spring constant when the current is 1.5A?

Solution:

For current I=1.5A, θ=55.5 degree=0.96865 rad

$$\frac{dL}{d\theta} = \frac{577.65 - 576.5}{60 - 49.5} = 0.11\mu H/\deg ree = 6.3\mu H/rad$$

Deflecting torque , $T_d = \frac{1}{2} I^2 \frac{dL}{d\theta} = \frac{1}{2}(1.5)^2 \times 6.3 \times 10^{-6} = 7.09 \times 10^{-6} N - m$

Spring constant, $K = \frac{T_d}{\theta} = \frac{7.09 \times 10^{-6}}{0.968} = 7.319 \times 10^{-6} \frac{N - m}{rad}$

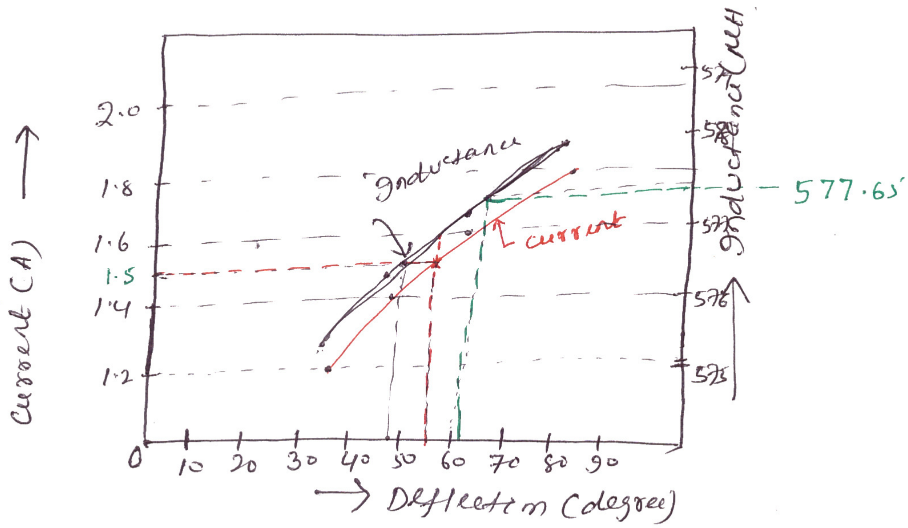


Fig. 1.25

Example: 1.9

For a certain dynamometer ammeter the mutual inductance 'M' varies with deflection θ as $M = -6\cos(\theta + 30^\circ) mH$ .Find the deflecting torque produced by a direct current of 50mA corresponding to a deflection of 60^0.

Solution:

$$T_d = I_1 I_2 \frac{dM}{d\theta} = I^2 \frac{dM}{d\theta}$$

$$M = -6\cos(\theta + 30^\circ)$$

$$\frac{dM}{d\theta} = 6\sin(\theta + 30) mH$$

$$\frac{dM}{d\theta}\Big|_{\theta=60} = 6\sin 90 = 6mH/\deg$$

$$T_d = I^2 \frac{dM}{d\theta} = (50 \times 10^{-3})^2 \times 6 \times 10^{-3} = 15 \times 10^{-6} N - m$$

Example: 1.10

The inductance of a moving iron ammeter with a full scale deflection of 90^0 at 1.5A, is given by the expression $L = 200 + 40\theta - 4\theta^2 - \theta^3 \mu H$, where θ is deflection in radian from the zero position. Estimate the angular deflection of the pointer for a current of 1.0A.

Solution:

$$L = 200 + 40\theta - 4\theta^2 - \theta^3 \mu H$$

$$\left.\frac{dL}{d\theta}\right|_{\theta=90^\circ} = 40 - 8\theta - 3\theta^2 \mu H / rad$$

$$\left.\frac{dL}{d\theta}\right|_{\theta=90^\circ} = 40 - 8\times\frac{\Pi}{2} - 3(\frac{\Pi}{2})^2 \mu H / rad = 20\mu H / rad$$

$$\therefore \theta = \frac{1}{2K} I^2 \left(\frac{dL}{d\theta}\right)$$

$$\frac{\Pi}{2} = \frac{1}{2}\frac{(1.5)^2}{K} \times 20\times10^{-6}$$

K=Spring constant=$14.32\times10^{-6} N - m / rad$

For I=1A, $\therefore \theta = \frac{1}{2K} I^2 \left(\frac{dL}{d\theta}\right)$

$$\therefore \theta = \frac{1}{2}\times\frac{(1)^2}{14.32\times10^{-6}}\left(40 - 8\theta - 3\theta^2\right)$$

$$3\theta + 36.64\theta^2 - 40 = 0$$

$$\theta = 1.008 rad, 57.8^\circ$$

Example: 1.11

The inductance of a moving iron instrument is given by $L = 10 + 5\theta - \theta^2 - \theta^3 \mu H$, where θ is the deflection in radian from zero position. The spring constant is $12\times10^{-6} N - m / rad$. Estimate the deflection for a current of 5A.

Solution:

$$\frac{dL}{d\theta} = (5 - 2\theta)\frac{\mu H}{rad}$$

$$\therefore \theta = \frac{1}{2K} I^2 \left(\frac{dL}{d\theta}\right)$$

$$\therefore \theta = \frac{1}{2} \times \frac{(5)^2}{12 \times 10^{-6}} (5 - 2\theta) \times 10^{-6}$$

$$\therefore \theta = 1.69 rad, 96.8^{\circ}$$

Example: 1.12

The following figure gives the relation between deflection and inductance of a moving iron instrument.

Deflection (degree)	20	30	40	50	60	70	80	90
Inductance (μH)	335	345	355.5	366.5	376.5	385	391.2	396.5

Find the current and the torque to give a deflection of (a) 30^0 (b) 80^0 . Given that control spring constant is $0.4 \times 10^{-6} N - m / \deg ree$

Solution:

$$\theta = \frac{1}{2K} I^2 \left(\frac{dL}{d\theta}\right)$$

(a) For $\theta = 30^{\circ}$

The curve is linear

$$\therefore \left(\frac{dL}{d\theta}\right)_{\theta=30} = \frac{355.5 - 335}{40 - 20} = 1.075 \mu H / \deg ree = 58.7 \mu H / rad$$

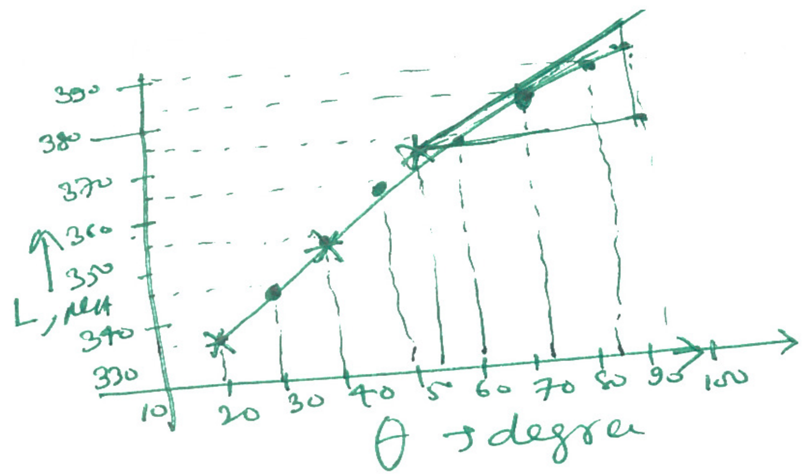


Fig. 1.26

Example: 1.13

In an electrostatic voltmeter the full scale deflection is obtained when the moving plate turns through 90^0. The torsional constant is $10\times10^{-6}\,N-m/rad$. The relation between the angle of deflection and capacitance between the fixed and moving plates is given by

Deflection (degree)	0	10	20	30	40	50	60	70	80	90
Capacitance (PF)	81.4	121	156	189.2	220	246	272	294	316	334

Find the voltage applied to the instrument when the deflection is 90^0?

Solution:

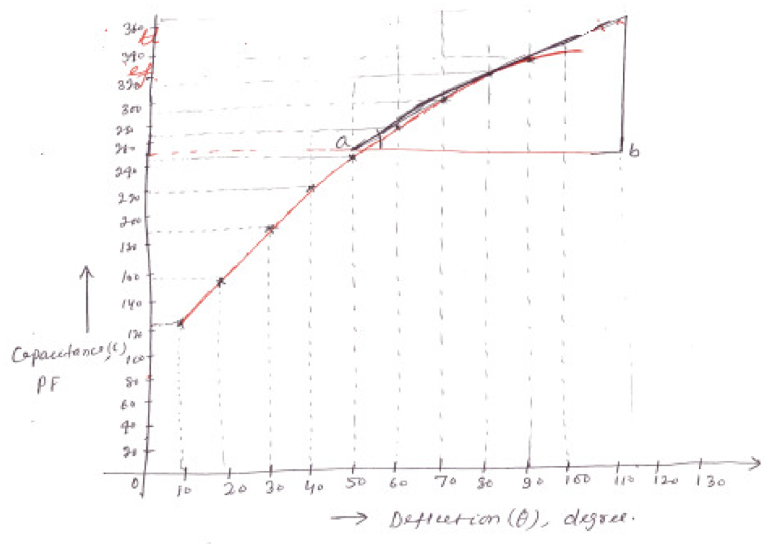


Fig. 1.27

$$\frac{dC}{d\theta} = \tan\theta = \frac{bc}{ab} = \frac{370-250}{110-44} = 1.82PF/\deg ree = 104.2PF/rad$$

Spring constant $K = 10\times10^{-6}\frac{N-m}{rad} = 0.1745\times10^{-6}N-m/\deg ree$

$$\theta = \frac{1}{2K}V^2\left(\frac{dC}{d\theta}\right) \Rightarrow V = \sqrt{\frac{2K\theta}{\frac{dC}{d\theta}}}$$

$$V = \sqrt{\frac{2\times0.1745\times10^{-6}\times90}{104.2\times10^{-12}}} = 549volt$$

Example: 1.14

Design a multi range d.c. mille ammeter using a basic movement with an internal resistance $R_m = 50\Omega$ and a full scale deflection current $I_m = 1mA$. The ranges required are 0-10mA; 0-50mA; 0-100mA and 0-500mA.

Solution:

Case-I 0-10mA

Multiplying power $m = \frac{I}{I_m} = \frac{10}{1} = 10$

$\therefore$ Shunt resistance $R_{sh1} = \frac{R_m}{m-1} = \frac{50}{10-1} = 5.55\Omega$

Case-II 0-50mA

$$m = \frac{50}{1} = 50$$

$$R_{sh2} = \frac{R_m}{m-1} = \frac{50}{50-1} = 1.03\Omega$$

Case-III 0-100mA, $m = \frac{100}{1} = 100\Omega$

$$R_{sh3} = \frac{R_m}{m-1} = \frac{50}{100-1} = 0.506\Omega$$

Case-IV 0-500mA**,** $m = \frac{500}{1} = 500\Omega$

$$R_{sh4} = \frac{R_m}{m-1} = \frac{50}{500-1} = 0.1\Omega$$

Example: 1.15

A moving coil voltmeter with a resistance of 20Ω gives a full scale deflection of 120^0, when a potential difference of 100mV is applied across it. The moving coil has dimension of 30mm*25mm and is wounded with 100 turns. The control spring constant is $0.375\times10^{-6} N-m/\deg ree.$ Find the flux density, in the air gap. Find also the diameter of copper wire of coil winding if 30% of instrument resistance is due to coil winding. The specific resistance for copper=$1.7\times10^{-8}\Omega m$.

Solution:

Data given

$V_m = 100mV$

$R_m = 20\Omega$

$\theta = 120^{\circ}$

N=100

$K = 0.375\times10^{-6} N-m/\deg ree$

$R_C = 30\% of R_m$

$\rho = 1.7\times10^{-8}\Omega m$

$$I_m = \frac{V_m}{R_m} = 5\times10^{-3} A$$

$T_d = BANI, T_C = K\theta = 0.375\times10^{-6}\times120 = 45\times10^{-6} N-m$

$$B = \frac{T_d}{ANI} = \frac{45\times10^{-6}}{30\times25\times10^{-6}\times100\times5\times10^{-3}} = 0.12 wb/m^2$$

$R_C = 0.3\times20 = 6\Omega$

Length of mean turn path =2(a+b) =2(55)=110mm

$$R_C = N\left(\frac{\rho l}{A}\right)$$

$$A = \frac{N\times\rho\times(l_t)}{R_C} = \frac{100\times1.7\times10^{-8}\times110\times10^{-3}}{6}$$

$= 3.116 \times 10^{-8} m^2$

$= 31.16 \times 10^{-3} mm^2$

$A = \frac{\Pi}{4} d^2 \Rightarrow d = 0.2mm$

Example: 1.16

A moving coil instrument gives a full scale deflection of 10mA, when the potential difference across its terminal is 100mV. Calculate

(1) The shunt resistance for a full scale deflection corresponding to 100A

(2) The resistance for full scale reading with 1000V.

Calculate the power dissipation in each case?

Solution:

Data given

$I_m = 10mA$

$V_m = 100mV$

$I = 100A$

$$I = I_m \left(1 + \frac{R_m}{R_{sh}}\right)$$

$$100 = 10 \times 10^{-3} \left(1 + \frac{10}{R_{sh}}\right)$$

$R_{sh} = 1.001 \times 10^{-3} \Omega$

$R_{se} = ??, V = 1000V$

$$R_m = \frac{V_m}{I_m} = \frac{100}{10} = 10\Omega$$

$$V = V_m \left(1 + \frac{R_{se}}{R_m}\right)$$

$$1000 = 100 \times 10^{-3} \left(1 + \frac{R_{se}}{10}\right)$$

$\therefore R_{se} = 99.99 K\Omega$

Example: 1.17

Design an Aryton shunt to provide an ammeter with current ranges of 1A,5A,10A and 20A. A basic meter with an internal resistance of 50w and a full scale deflection current of 1mA is to be used.

Solution: Data given

$$I_m = 1\times10^{-3}A \quad R_m = 50\Omega \quad \left| \begin{array}{l} I_1 = 1A \\ I_2 = 5A \\ I_3 = 10A \\ I_4 = 20A \end{array} \right| \begin{array}{l} m_1 = \dfrac{I_1}{I_m} = 1000A \\ m_2 = \dfrac{I_2}{I_m} = 5000A \\ m_3 = \dfrac{I_3}{I_m} = 10000A \\ m_4 = \dfrac{I_4}{I_m} = 20000A \end{array}$$

$$R_{sh1} = \frac{R_m}{m_1 - 1} = \frac{50}{1000-1} = 0.05\Omega$$
$$R_{sh2} = \frac{R_m}{m_2 - 1} = \frac{50}{5000-1} = 0.01\Omega$$
$$R_{sh3} = \frac{R_m}{m_3 - 1} = \frac{50}{10000-1} = 0.005\Omega$$
$$R_{sh4} = \frac{R_m}{m_4 - 1} = \frac{50}{20000-1} = 0.0025\Omega$$

$\therefore$ The resistances of the various section of the universal shunt are

$$R_1 = R_{sh1} - R_{sh2} = 0.05 - 0.01 = 0.04\Omega$$
$$R_2 = R_{sh2} - R_{sh3} = 0.01 - 0.005 = 0.005\Omega$$
$$R_3 = R_{sh3} - R_{sh4} = 0.005 - 0.025 = 0.0025\Omega$$
$$R_4 = R_{sh4} = 0.0025\Omega$$

Example: 1.18

A basic d' Arsonval meter movement with an internal resistance $R_m = 100\Omega$ and a full scale current of $I_m = 1mA$ is to be converted in to a multi range d.c. voltmeter with ranges of 0-10V, 0-50V, 0-250V, 0-500V. Find the values of various resistances using the potential divider arrangement.

Solution:

Data given

$$m_1 = \frac{V_1}{V_m} = \frac{10}{100\times10^{-3}} = 100$$

$$R_m = 100\Omega$$

$$I_m = 1mA$$

$$m_2 = \frac{V_2}{V_m} = \frac{50}{100\times10^{-3}} = 500$$

$$V_m = I_m \times R_m$$

$$V_m = 100\times1\times10^{-3}$$

$$m_3 = \frac{V_3}{V_m} = \frac{250}{100\times10^{-3}} = 2500$$

$$V_m = 100mV$$

$$m_4 = \frac{V_4}{V_m} = \frac{500}{100\times10^{-3}} = 5000$$

$$R_1 = (m_1 - 1)R_m = (100-1)\times100 = 9900\Omega$$

$$R_2 = (m_2 - m_1)R_m = (500-100)\times100 = 40K\Omega$$

$$R_3 = (m_3 - m_2)R_m = (2500-500)\times100 = 200K\Omega$$

$$R_4 = (m_4 - m_3)R_m = (5000-2500)\times100 = 250K\Omega$$

AC BRIDGES

2.1 General form of A.C. bridge

AC bridge are similar to D.C. bridge in topology(way of connecting).It consists of four arm AB,BC,CD and DA .Generally the impedance to be measured is connected between ‘A’ and ‘B’. A detector is connected between ‘B’ and ’D’. The detector is used as null deflection instrument. Some of the arms are variable element. By varying these elements, the potential values at ‘B’ and ‘D’ can be made equal. This is called balancing of the bridge.

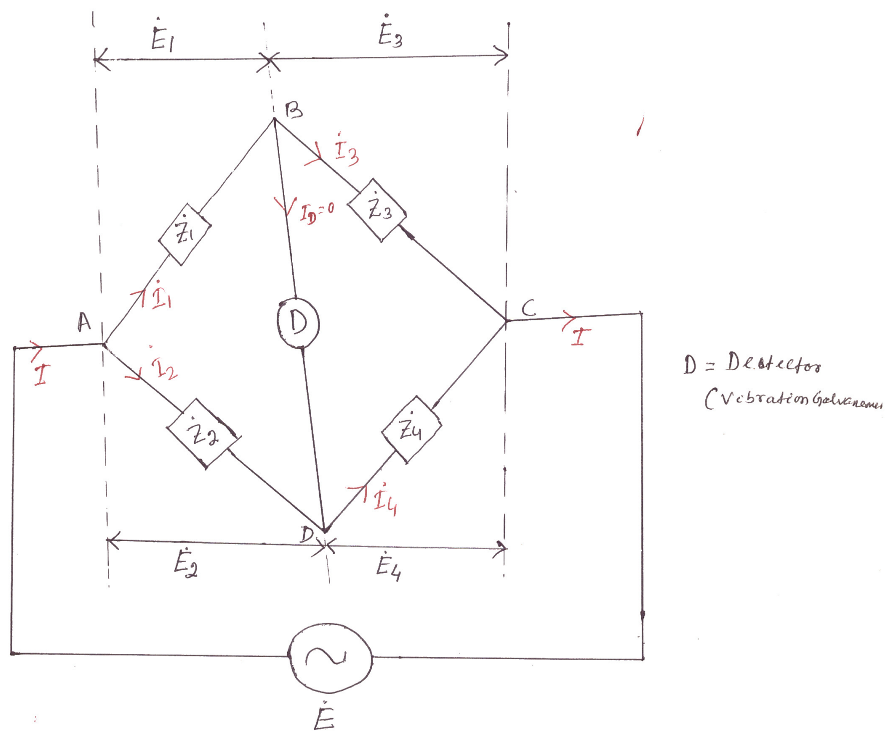


Fig. 2.1 General form of A.C. bridge

At the balance condition, the current through detector is zero.

$$\therefore \dot{I}_1 = \dot{I}_3$$

$$\dot{I}_2 = \dot{I}_4$$

$$\therefore \frac{\dot{I}_1}{\dot{I}_2} = \frac{\dot{I}_3}{\dot{I}_4} \tag{2.1}$$

At balance condition,

Voltage drop across 'AB'=voltage drop across 'AD'.

$$\dot{E}_1 = \dot{E}_2$$

$$\therefore \dot{I}_1 \dot{Z}_1 = \dot{I}_2 \dot{Z}_2 \tag{2.2}$$

Similarly, Voltage drop across 'BC'=voltage drop across 'DC'

$$\dot{E}_3 = \dot{E}_4$$

$$\therefore \dot{I}_3 \dot{Z}_3 = \dot{I}_4 \dot{Z}_4 \tag{2.3}$$

From Eqn. (2.2), we have $\therefore \frac{\dot{I}_1}{\dot{I}_2} = \frac{\dot{Z}_2}{\dot{Z}_1}$ (2.4)

From Eqn. (2.3), we have $\therefore \frac{\dot{I}_3}{\dot{I}_4} = \frac{\dot{Z}_4}{\dot{Z}_3}$ (2.5)

From equation -2.1, it can be seen that, equation -2.4 and equation-2.5 are equal.

$$\therefore \frac{\dot{Z}_2}{\dot{Z}_1} = \frac{\dot{Z}_4}{\dot{Z}_3}$$

$$\therefore \dot{Z}_1 \dot{Z}_4 = \dot{Z}_2 \dot{Z}_3$$

Products of impedances of opposite arms are equal.

$$\therefore |Z_1|\angle\theta_1 |Z_4|\angle\theta_4 = |Z_2|\angle\theta_2 |Z_3|\angle\theta_3$$

$$\Rightarrow |Z_1||Z_4|\angle\theta_1 + \theta_4 = |Z_2||Z_3|\angle\theta_2 + \theta_3$$

$$|Z_1||Z_4| = |Z_2||Z_3|$$

$$\theta_1 + \theta_4 = \theta_2 + \theta_3$$

* For balance condition, magnitude on either side must be equal.

* Angle on either side must be equal.

Summary

For balance condition,

- $\dot{I}_1 = \dot{I}_3, \quad \dot{I}_2 = \dot{I}_4$
- $|Z_1||Z_4| = |Z_2||Z_3|$
- $\theta_1 + \theta_4 = \theta_2 + \theta_3$
- $\dot{E}_1 = \dot{E}_2$ & $\dot{E}_3 = \dot{E}_4$

2.2 Types of detector

The following types of instruments are used as detector in A.C. bridge.

- Vibration galvanometer
- Head phones (speaker)
- Tuned amplifier

2.2.1 Vibration galvanometer

Between the point 'B' and 'D' a vibration galvanometer is connected to indicate the bridge balance condition. This A.C. galvanometer which works on the principle of resonance. The A.C. galvanometer shows a dot, if the bridge is unbalanced.

2.2.2 Head phones

Two speakers are connected in parallel in this system. If the bridge is unbalanced, the speaker produced more sound energy. If the bridge is balanced, the speaker do not produced any sound energy.

2.2.3 Tuned amplifier

If the bridge is unbalanced the output of tuned amplifier is high. If the bridge is balanced, output of amplifier is zero.

2.3 Measurements of inductance

2.3.1 Maxwell's inductance bridge

The choke for which R_1 and L_1 have to measure connected between the points 'A' and 'B'. In this method the unknown inductance is measured by comparing it with the standard inductance.

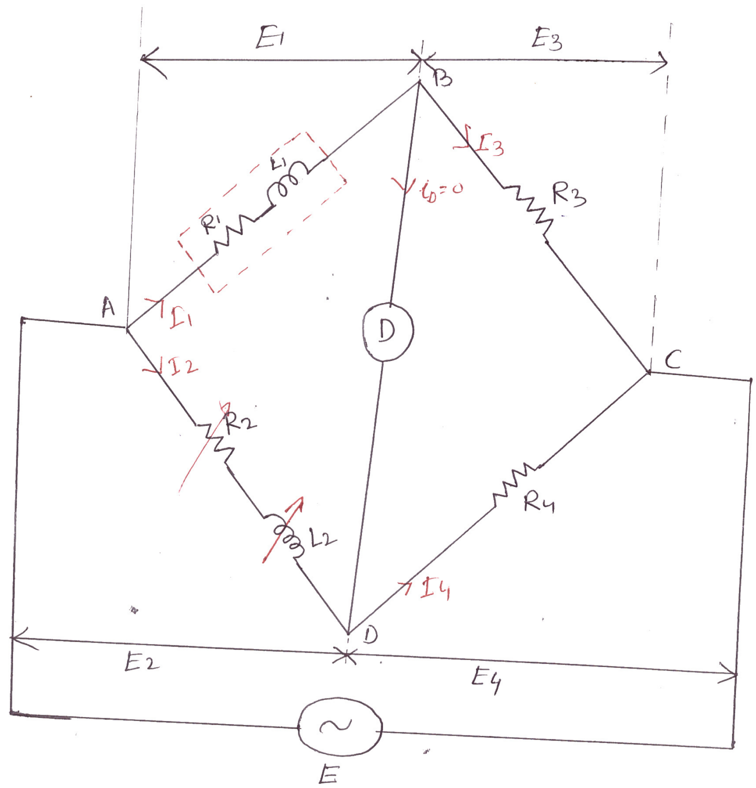


Fig. 2.2 Maxwell's inductance bridge

L_2 is adjusted, until the detector indicates zero current.

Let R_1= unknown resistance

L_1= unknown inductance of the choke.

L_2= known standard inductance

R_1,R_2,R_4= known resistances.

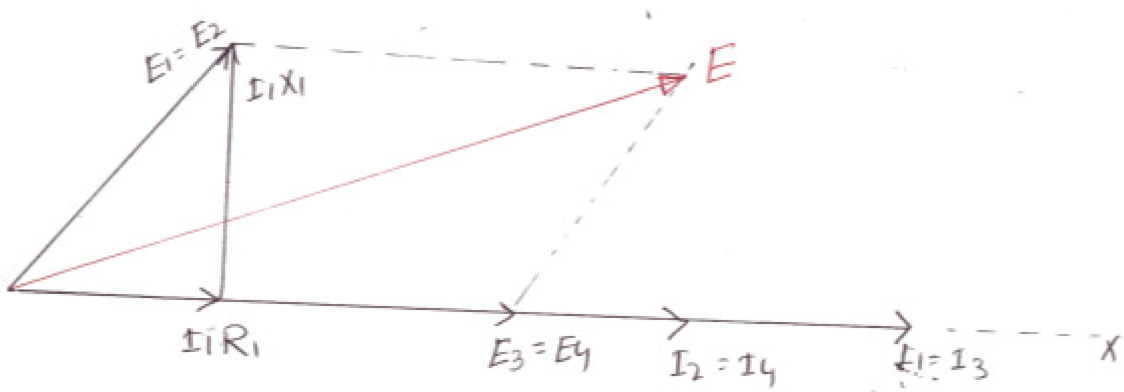


Fig 2.3 Phasor diagram of Maxwell's inductance bridge

At balance condition, $\dot{Z}_1 \dot{Z}_4 = \dot{Z}_2 \dot{Z}_3$

$$(R_1 + jXL_1)R_4 = (R_2 + jXL_2)R_3$$

$$(R_1 + jwL_1)R_4 = (R_2 + jwL_2)R_3$$

$$R_1R_4 + jwL_1R_4 = R_2R_3 + jwL_2R_3$$

Comparing real part,

$$R_1R_4 = R_2R_3$$

$$\therefore R_1 = \frac{R_2R_3}{R_4} \tag{2.6}$$

Comparing the imaginary parts,

$$wL_1R_4 = wL_2R_3$$

$$L_1 = \frac{L_2R_3}{R_4} \tag{2.7}$$

Q-factor of choke, $Q = \dfrac{WL_1}{R_1} = \dfrac{WL_2R_3R_4}{R_4R_2R_3}$

$$Q = \frac{WL_2}{R_2} \tag{2.8}$$

Advantages

- ✓ Expression for R_1 and L_1 are simple.
- ✓ Equations area simple
- ✓ They do not depend on the frequency (as w is cancelled)
- ✓ R_1 and L_1 are independent of each other.

Disadvantages

- ✓ Variable inductor is costly.
- ✓ Variable inductor is bulky.

2.3.2 Maxwell's inductance capacitance bridge

Unknown inductance is measured by comparing it with standard capacitance. In this bridge, balance condition is achieved by varying 'C_4'.

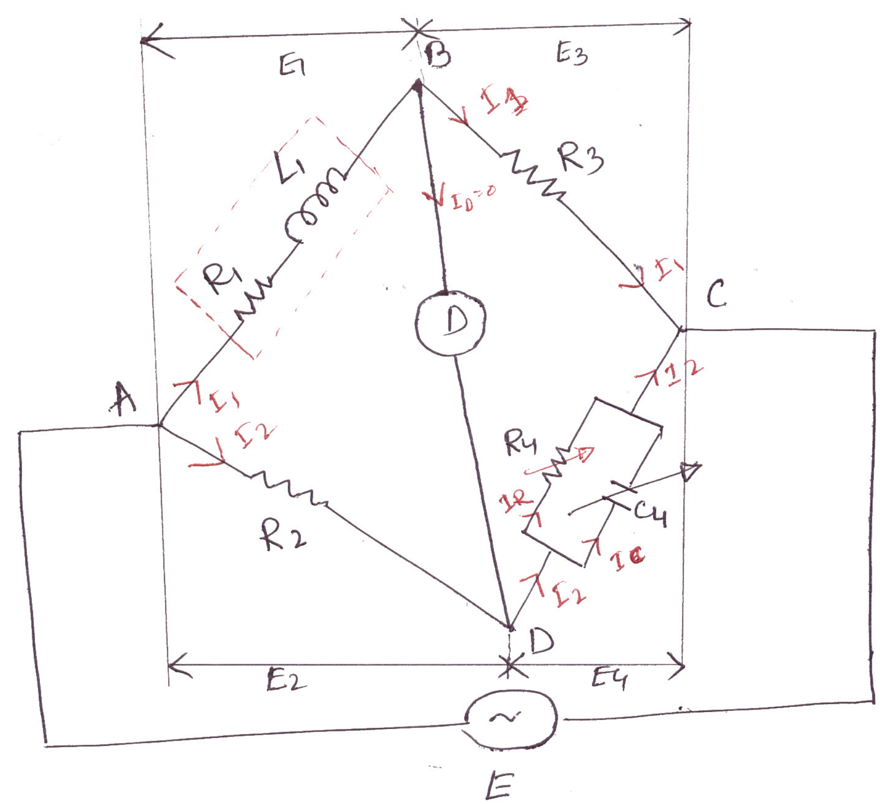


Fig 2.4 Maxwell's inductance capacitance bridge

At balance condition, $Z_1Z_4=Z_3Z_2$ (2.9)

$$Z_4 = R_4 \parallel \frac{1}{jwC_4} = \frac{R_4 \times \frac{1}{jwC_4}}{R_4 + \frac{1}{jwC_4}}$$

$$Z_4 = \frac{R_4}{jwR_4C_4 + 1} = \frac{R_4}{1 + jwR_4C_4} \tag{2.10}$$

$\therefore$ Substituting the value of Z_4 from eqn. (2.10) in eqn. (2.9) we get

$$(R_1 + jwL_1) \times \frac{R_4}{1 + jwR_4C_4} = R_2R_3$$

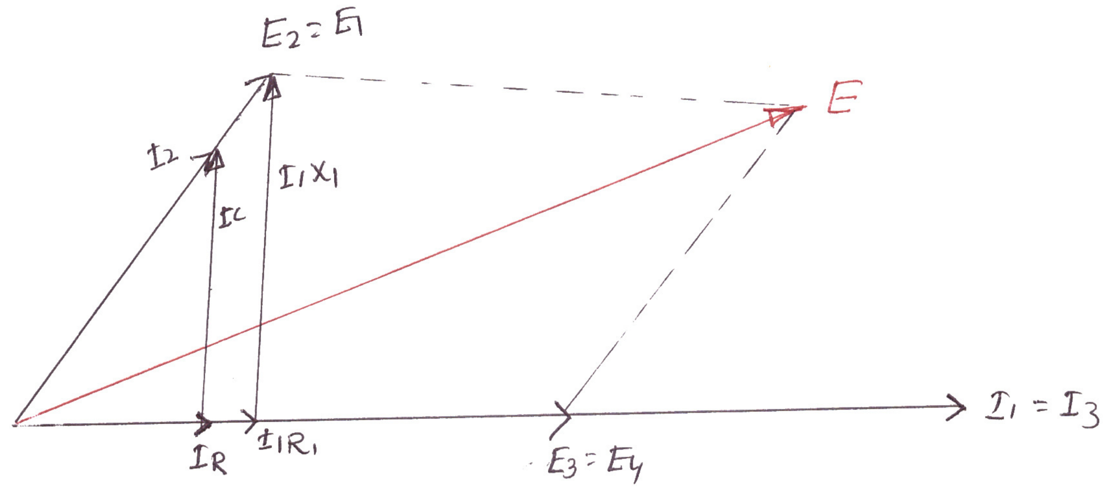


Fig 2.5 Phasor diagram of Maxwell's inductance capacitance bridge

$$(R_1 + jwL_1)R_4 = R_2R_3(1 + jwR_4C_4)$$

$$R_1R_4 + jwL_1R_4 = R_2R_3 + jwC_4R_4R_2R_3$$

Comparing real parts,

$$R_1R_4 = R_2R_3$$

$$\Rightarrow R_1 = \frac{R_2 R_3}{R_4} \tag{2.11}$$

Comparing imaginary part,

$$wL_1 R_4 = wC_4 R_4 R_2 R_3$$

$$L_1 = C_4 R_2 R_3 \tag{2.12}$$

Q-factor of choke,

$$Q = \frac{WL_1}{R_1} = w \times C_4 R_2 R_3 \times \frac{R_4}{R_2 R_3}$$

$$Q = wC_4 R_4 \tag{2.13}$$

Advantages

- ✓ Equation of L_1 and R_1 are simple.
- ✓ They are independent of frequency.
- ✓ They are independent of each other.
- ✓ Standard capacitor is much smaller in size than standard inductor.

Disadvantages

- ✓ Standard variable capacitance is costly.
- ✓ It can be used for measurements of Q-factor in the ranges of 1 to 10.
- ✓ It cannot be used for measurements of choke with Q-factors more than 10.

  We know that Q $=wC_4R_4$

For measuring chokes with higher value of Q-factor, the value of C_4 and R_4 should be higher. Higher values of standard resistance are very expensive. Therefore this bridge cannot be used for higher value of Q-factor measurements.

2.3.3 Hay's bridge

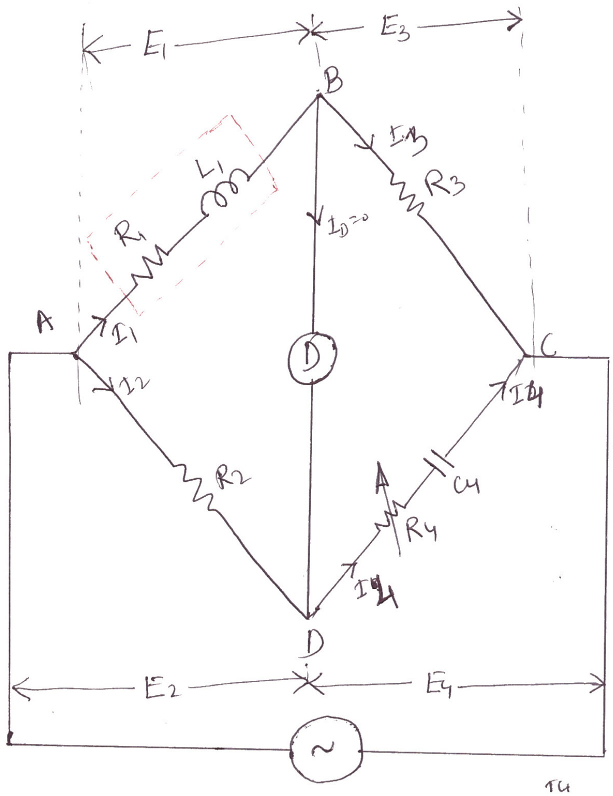


Fig 2.6 Hay's bridge

- $\dot{E}1 = I_1R_1 + jI_1X_1$
- $\dot{E} = \dot{E}1 + \dot{E}3$
- $\dot{E}4 = \dot{I}_4 R_4 + \dfrac{I_4}{jwC_4}$
- $\dot{E}3 = I_3R_3$

$$Z_4 = R_4 + \frac{1}{jwC_4} = \frac{1 + jwR_4C_4}{jwC_4}$$

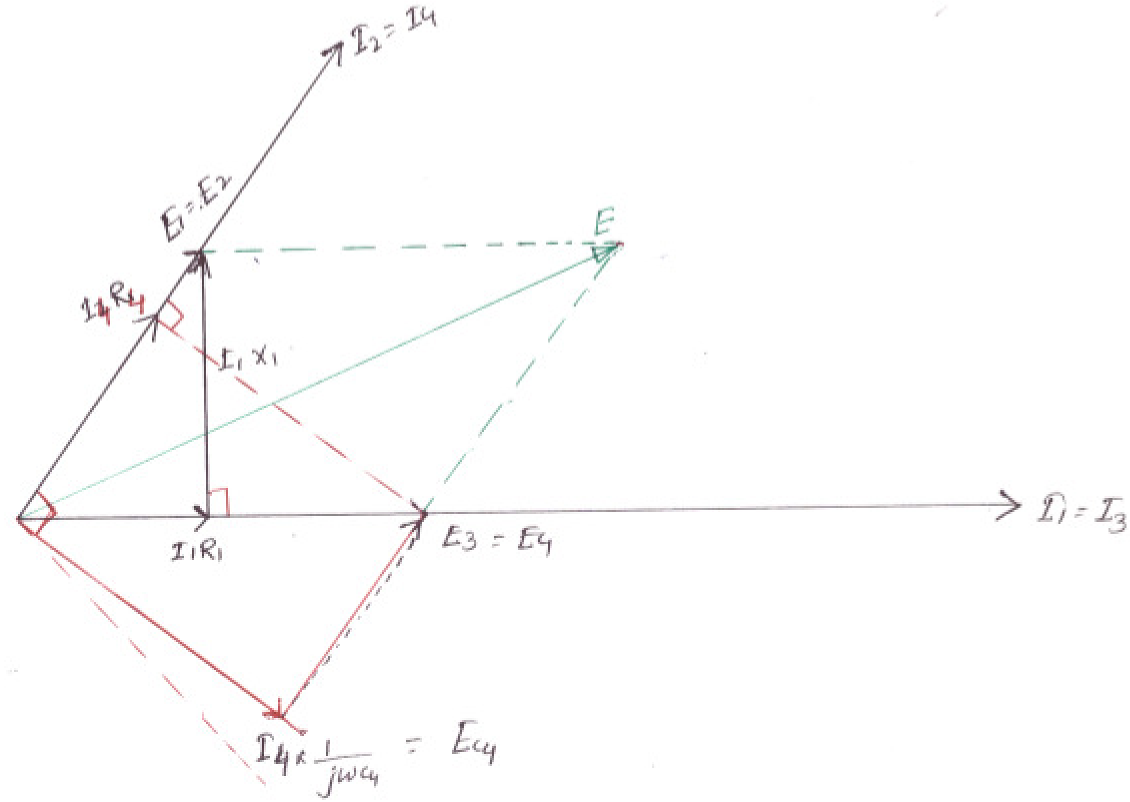


Fig 2.7 Phasor diagram of Hay's bridge

At balance condition, $Z_1Z_4=Z_3Z_2$

$$(R_1 + jwL_1)\left(\frac{1+ jwR_4C_4}{jwC_4}\right) = R_2R_3$$

$$(R_1 + jwL_1)(1+ jwR_4C_4) = jwR_2C_4R_3$$

$$R_1 + jwC_4R_4R_1 + jwL_1 + j^2w^2L_1C_4R_4 = jwC_4R_2R_3$$

$$(R_1 - w^2L_1C_4R_4) + j(wC_4R_4R_1 + wL_1) = jwC_4R_2R_3$$

Comparing the real term,

$$R_1 - w^2L_1C_4R_4 = 0$$

$$R_1 = w^2L_1C_4R_4 \tag{2.14}$$

Comparing the imaginary terms,

$$wC_4R_4R_1 + wL_1 = wC_4R_2R_3$$

$$C_4R_4R_1 + L_1 = C_4R_2R_3$$

$$L_1 = C_4R_2R_3 - C_4R_4R_1 \tag{2.15}$$

Substituting the value of R_1 fro eqn. 2.14 into eqn. 2.15, we have,

$$L_1 = C_4R_2R_3 - C_4R_4 \times w^2L_1C_4R_4$$

$$L_1 = C_4R_2R_3 - w^2L_1{C_4}^2{R_4}^2$$

$$L_1(1 + w^2L_1{C_4}^2{R_4}^2) = C_4R_2R_3$$

$$L_1 = \frac{C_4R_2R_3}{1 + w^2L_1{C_4}^2{R_4}^2} \tag{2.16}$$

Substituting the value of L_1 in eqn. 2.14 , we have

$$R_1 = \frac{w^2{C_4}^2R_2R_3R_4}{1 + w^2{C_4}^2{R_4}^2} \tag{2.17}$$

$$Q = \frac{wL_1}{R_1} = \frac{w \times C_4R_2R_3}{1 + w^2{C_4}^2{R_4}^2} \times \frac{1 + w^2{C_4}^2{R_4}^2}{w^2{C_4}^2R_4R_2R_3}$$

$$Q = \frac{1}{wC_4R_4} \tag{2.18}$$

Advantages

- ✓ Fixed capacitor is cheaper than variable capacitor.
- ✓ This bridge is best suitable for measuring high value of Q-factor.

Disadvantages

- ✓ Equations of L1and R1 are complicated.
- ✓ Measurements of R_1 and L_1 require the value of frequency.
- ✓ This bridge cannot be used for measuring low Q- factor.

2.3.4 Owen's bridge

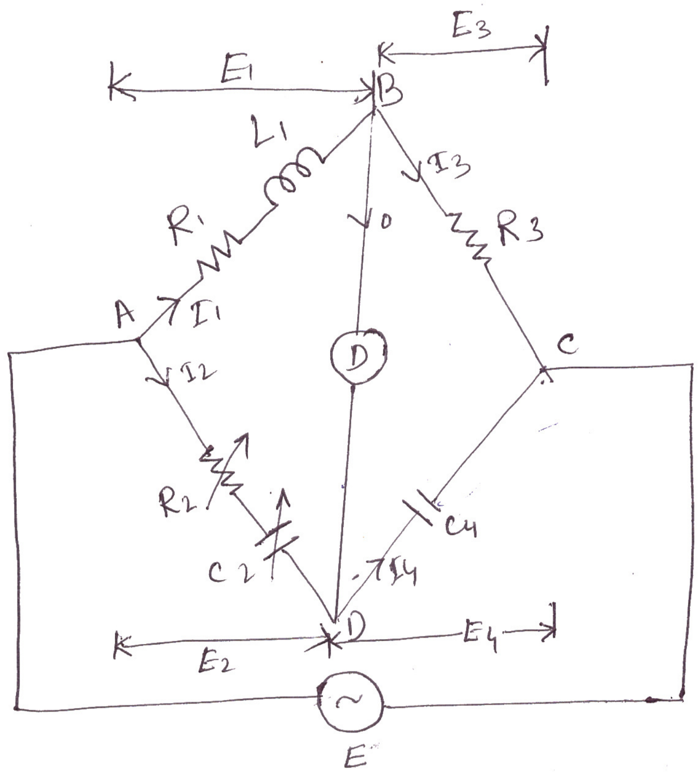


Fig 2.8 Owen's bridge

- ➢ $E_1 = I_1R_1 + jI_1X_1$
- ➢ I_4 leads E_4 by 90^0

- $\dot{E} = \dot{E}_1 + \dot{E}_3$
- $\dot{E}_2 = I_2 R_2 + \dfrac{I_2}{jwC_2}$

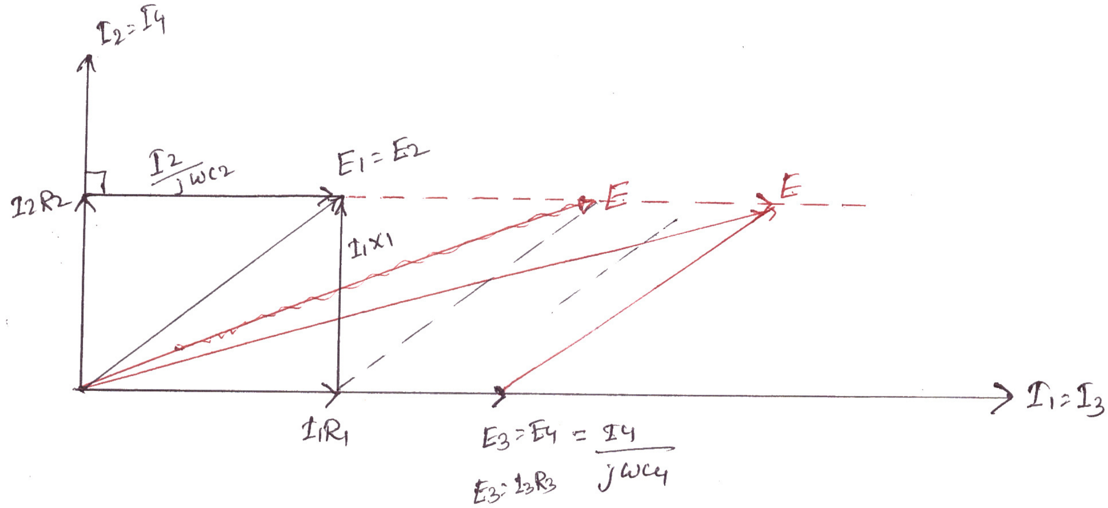


Fig 2.9 Phasor diagram of Owen's bridge

Balance condition, $\dot{Z}_1 \dot{Z}_4 = \dot{Z}_2 \dot{Z}_3$

$$Z_2 = R_2 + \frac{1}{jwC_2} = \frac{jwC_2R_2 + 1}{jwC_2}$$

$$\therefore (R_1 + jwL_1) \times \frac{1}{jwC_4} = \frac{(1 + jwR_2C_2) \times R_3}{jwC_2}$$

$$C_2(R_1 + jwL_1) = R_3C_4(1 + jwR_2C_2)$$

$$R_1C_2 + jwL_1C_2 = R_3C_4 + jwR_2C_2R_3C_4$$

Comparing real terms,

$$R_1C_2 = R_3C_4$$

$$R_1 = \frac{R_3 C_4}{C_2}$$

Comparing imaginary terms,

$$wL_1 C_2 = wR_2 C_2 R_3 C_4$$

$$L_1 = R_2 R_3 C_4$$

Q- factor = $\dfrac{WL_1}{R_1} = \dfrac{wR_2 R_3 C_4 C_2}{R_3 C_4}$

$$Q = wR_2 C_2$$

Advantages

- ✓ Expression for R_1 and L_1 are simple.
- ✓ R_1 and L_1 are independent of Frequency.

Disadvantages

- ✓ The Circuits used two capacitors.
- ✓ Variable capacitor is costly.
- ✓ Q-factor range is restricted.

2.3.5 Anderson's bridge

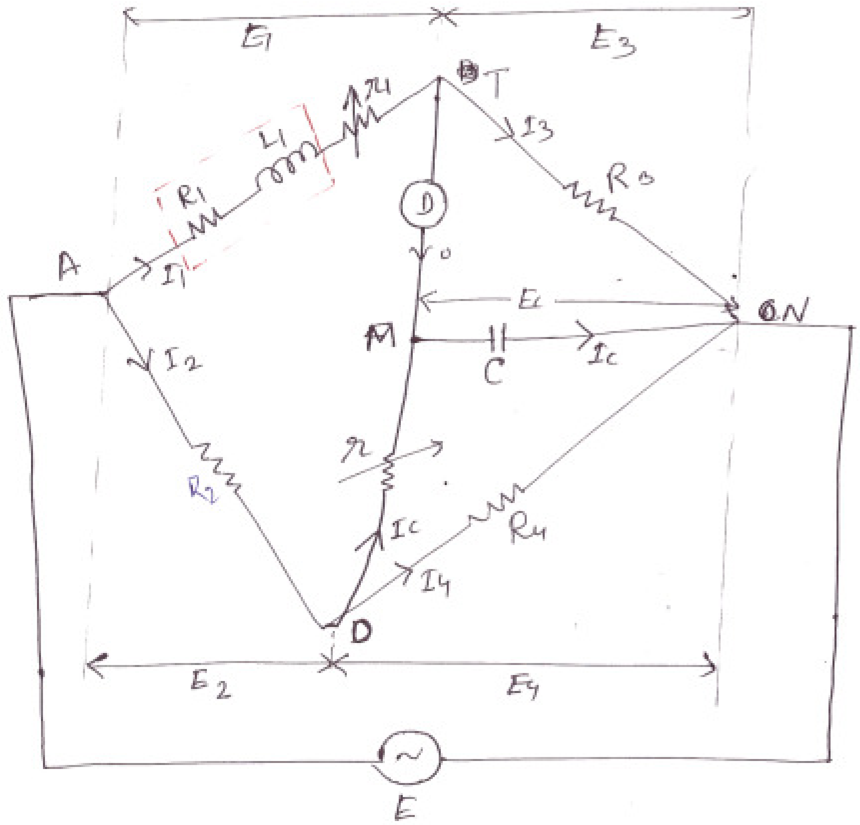


Fig 2.10 Anderson's bridge

- $\dot{E}_1 = I_1(R_1 + r_1) + jI_1X_1$
- $E_3 = E_C$
- $\dot{E}_4 = \dot{I}_C r + E_C$
- $I_2 = I_4 + I_C$
- $\bar{E}_2 + \bar{E}_4 = \bar{E}$
- $\bar{E}_1 + \bar{E}_3 = \bar{E}$

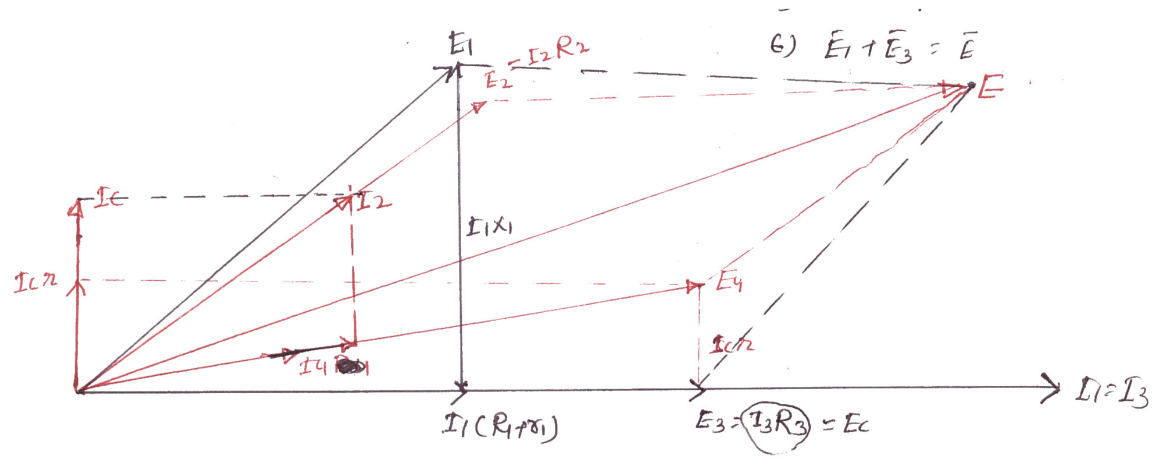


Fig 2.11 Phasor diagram of Anderson's bridge

Step-1 Take I_1 as references vector .Draw $I_1R_1^1$ in phase with I_1

$R_1^1 = (R_1 + r_1)$, I_1X_1 is $\perp_r$ to $I_1R_1^1$

$E_1 = I_1R_1^1 + jI_1X_1$

Step-2 $I_1 = I_3$, E_3 is in phase with I_3 , From the circuit ,

$E_3 = E_C$, I_C leads E_C by 90^0

Step-3 $E_4 = I_C r + E_C$

Step-4 Draw I_4 in phase with E_4 , By KCL , $\bar{I}_2 = \bar{I}_4 + \bar{I}_C$

Step-5 Draw E_2 in phase with I_2

Step-6 By KVL , $\bar{E}_1 + \bar{E}_3 = \bar{E}$ or $\bar{E}_2 + \bar{E}_4 = \bar{E}$

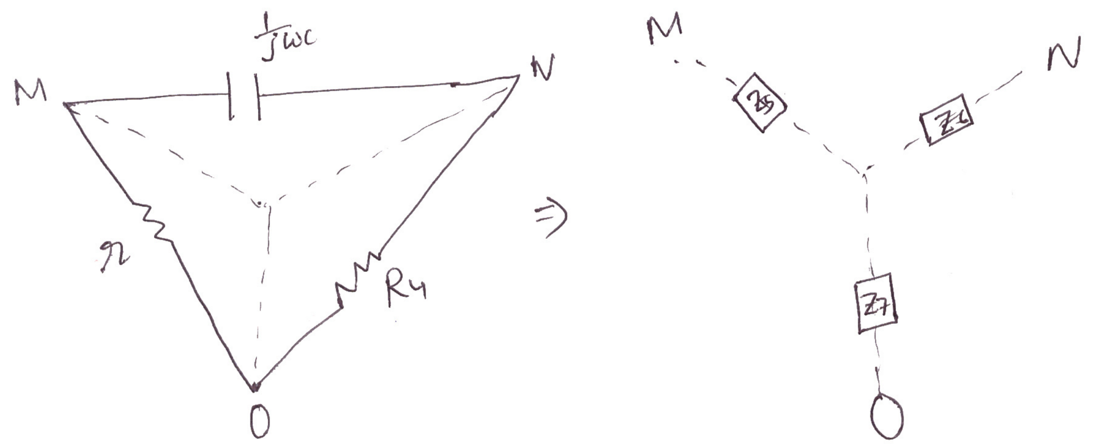


Fig 2.12 Equivalent delta to star conversion for the loop MON

$$Z_7 = \frac{R_4 \times r}{R_4 + r + \frac{1}{jwc}} = \frac{jwCR_4 r}{1 + jwC(R_4 + r)}$$

$$Z_6 = \frac{R_4 \times \frac{1}{jwC}}{R_4 + r + \frac{1}{jwc}} = \frac{R_4}{1 + jwC(R_4 + r)}$$

$$(R_1^1 + jwL_1) \times \frac{R_4}{1 + jwC(R_4 + r)} = R_3 (R_2 + \frac{jwCR_4 r}{1 + jwC(R_4 + r)})$$

$$\Rightarrow \frac{(R_1^1 + jwL_1) R_4}{1 + jwC(R_4 + r)} = R_3 \left[\frac{R_2 (1 + jwC(R_4 + r)) + jwCrR_4}{1 + jwC(R_4 + r)} \right]$$

$$\Rightarrow R_1^1 R_4 + jwL_1 R_4 = R_2 R_3 + jCwR_2 R_3 (r + R_4) + jwCrR_4 R_3$$

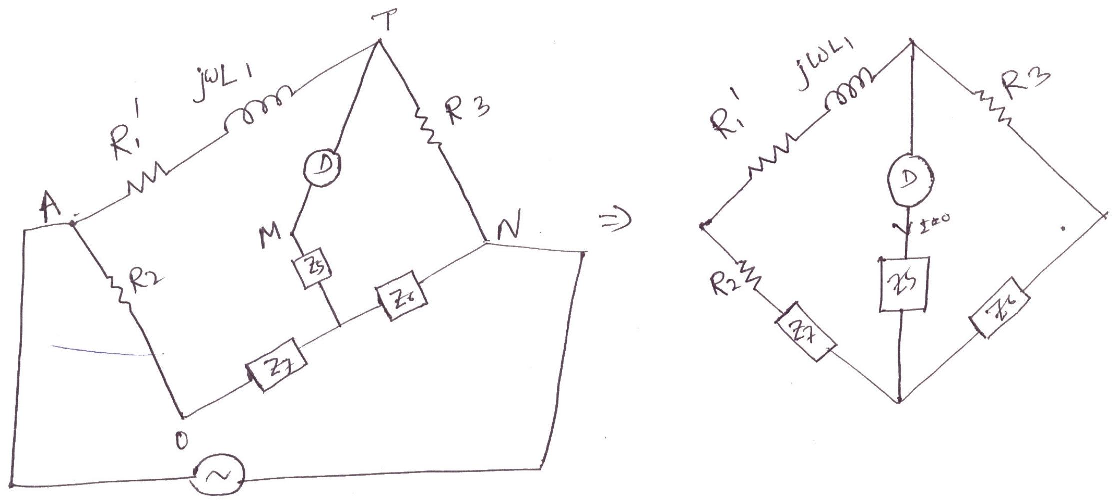


Fig 2.13 Simplified diagram of Anderson's bridge

Comparing real term,

$$R_1^1 R_4 = R_2 R_3$$

$$(R_1 + r_1) R_4 = R_2 R_3$$

$$R_1 = \frac{R_2 R_3}{R_4} - r_1$$

Comparing the imaginary term,

$$wL_1 R_4 = wCR_2 R_3 (r + R_4) + wcrR_3 R_4$$

$$L_1 = \frac{R_2 R_3 C}{R_4}(r + R_4) + R_3 rC$$

$$L_1 = R_3 C \left[\frac{R_2}{R_4}(r + R_4) + r \right]$$

Advantages

- ✓ Variable capacitor is not required.
- ✓ Inductance can be measured accurately.
- ✓ R_1 and L_1 are independent of frequency.
- ✓ Accuracy is better than other bridges.

Disadvantages

- ✓ Expression for R_1 and L_1 are complicated.
- ✓ This is not in the standard form A.C. bridge.

2.4 Measurement of capacitance and loss angle. (Dissipation factor)

2.4.1 Dissipation factors (D)

A practical capacitor is represented as the series combination of small resistance and ideal capacitance.

From the vector diagram, it can be seen that the angle between voltage and current is slightly less than 90^0. The angle 'δ' is called loss angle.

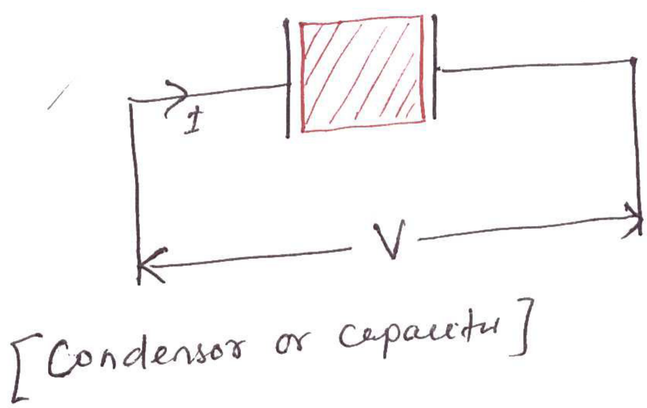


Fig 2.14 Condensor or capacitor

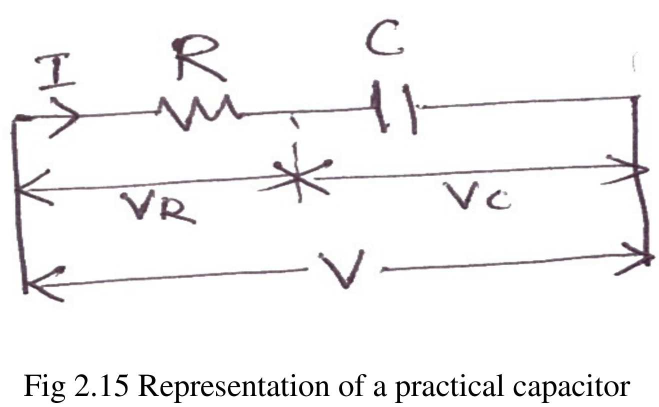


Fig 2.15 Representation of a practical capacitor

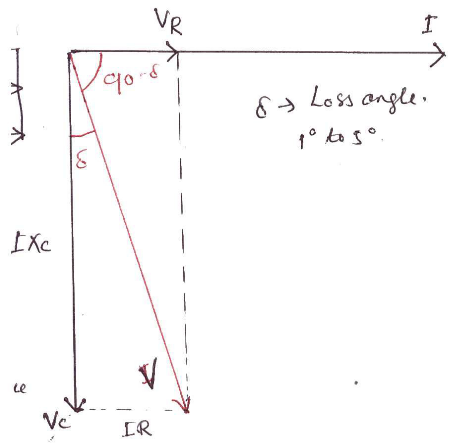


Fig 2.16 Vector diagram for a practical capacitor

A dissipation factor is defined as 'tan δ'.

$$\therefore \tan\delta = \frac{IR}{IX_C} = \frac{R}{X_C} = wCR$$

$$D = wCR$$

$$D = \frac{1}{Q}$$

$$D = \tan\delta = \frac{\sin\delta}{\cos\delta} \cong \frac{\delta}{1}$$ For small value of 'δ' in radians

$D \cong \delta \cong$ Loss Angle ('δ' must be in radian)

2.4.2 Desauty's Bridge

C_1= Unknown capacitance

At balance condition,

$$\frac{1}{jwC_1} \times R_4 = \frac{1}{jwC_2} \times R_3$$

$$\frac{R_4}{C_1} = \frac{R_3}{C_2}$$

$$\Rightarrow C_1 = \frac{R_4 C_2}{R_3}$$

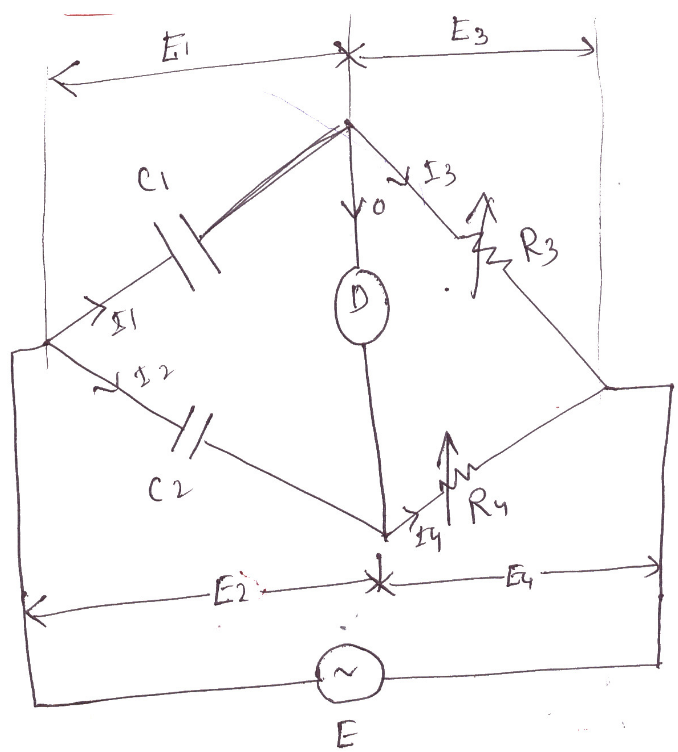


Fig 2.17 Desauty's bridge

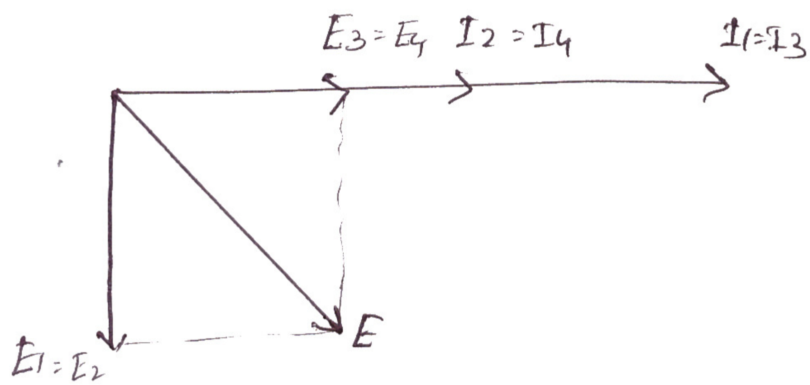


Fig 2.18 Phasor diagram of Desauty's bridge

2.4.3 Modified desauty's bridge

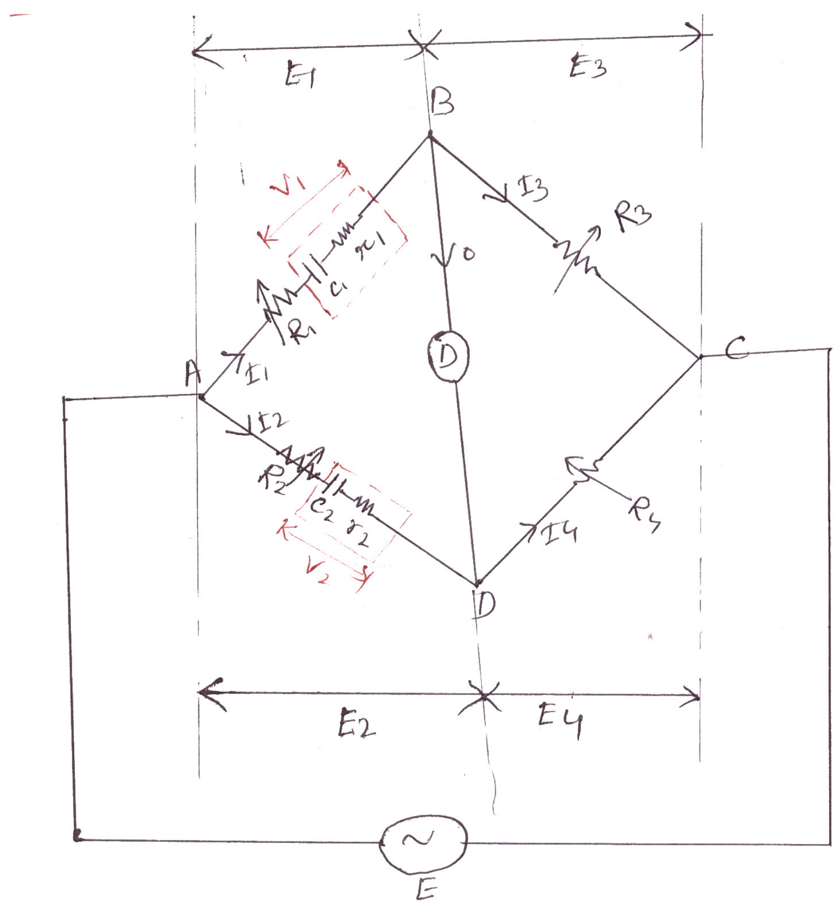


Fig 2.19 Modified Desauty's bridge

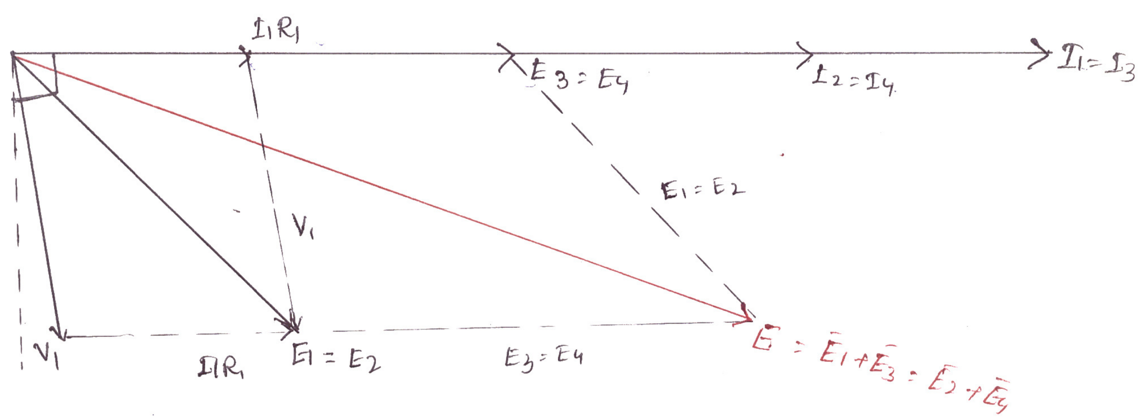


Fig 2.20 Phasor diagram of Modified Desauty's bridge

$R_1^1 = (R_1 + r_1)$

$R_2^1 = (R_2 + r_2)$

At balance condition, $(R_1^1 + \frac{1}{jwC_1})R_4 = R_3(R_2^1 + \frac{1}{jwC_2})$

$R_1^1 R_4 + \frac{R_4}{jwC_1} = R_3 R_2^1 + \frac{R_3}{jwC_2})$

Comparing the real term, $R_1^1 R_4 = R_3 R_2^1$

$R_1^1 = \frac{R_3 R_2^1}{R_4}$

$R + r_1 = \frac{(R_2 + r_2)R_3}{R_4}$

Comparing imaginary term,

$\frac{R_4}{wC_1} = \frac{R_3}{wC_2}$

$C_1 = \frac{R_4 C_2}{R_3}$

Dissipation factor $D = wC_1 r_1$

Advantages

- ✓ r_1 and c_1 are independent of frequency.
- ✓ They are independent of each other.
- ✓ Source need not be pure sine wave.

2.4.4 Schering bridge

$E_1 = I_1 r_1 - jI_1 X_4$

$C_2 = C_4$= Standard capacitor (Internal resistance=0)

C_4= Variable capacitance.

C_1= Unknown capacitance.

r_1= Unknown series equivalent resistance of the capacitor.

R_3=R_4= Known resistor.

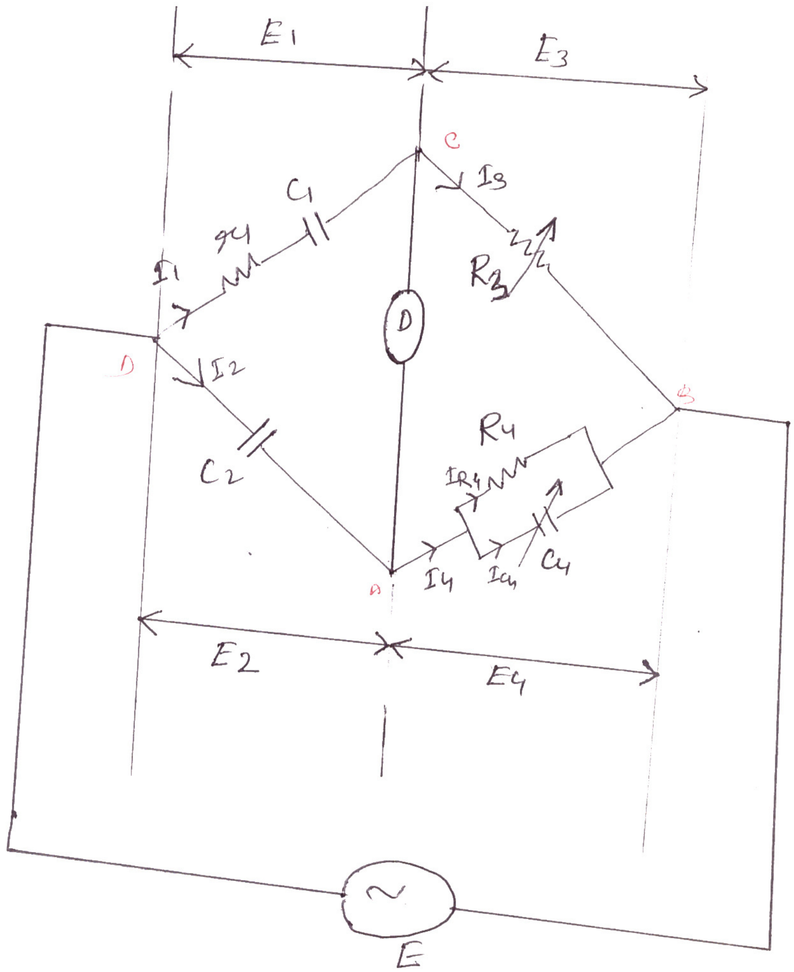


Fig 2.21 Schering bridge

$$Z_1 = r_1 + \frac{1}{jwC_1} = \frac{jwC_1 r_1 + 1}{jwC_1}$$

$$Z_4 = \frac{R_4 \times \dfrac{1}{jwC_4}}{R_4 + \dfrac{1}{jwC_4}} = \frac{R_4}{1 + jwC_4 R_4}$$

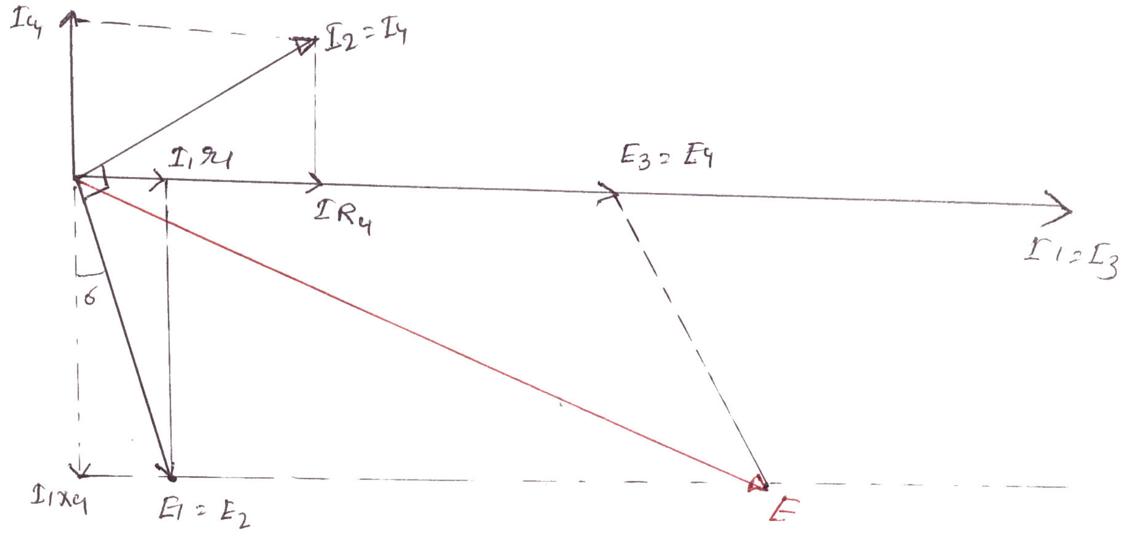


Fig 2.22 Phasor diagram of Schering bridge

At balance condition, $\dot{Z}_1 \dot{Z}_4 = \dot{Z}_2 \dot{Z}_3$

$$\frac{1+jwC_1r_1}{jwC_1} \times \frac{R_4}{1+jwC_4R_4} = \frac{R_3}{jwC_2}$$

$$(1+jwC_1r_1)R_4C_2 = R_3C_1(1+jwC_4r_4)$$

$$R_2C_2 + jwC_1r_1R_4C_2 = R_3C_1 + jwC_4R_4R_3C_1$$

Comparing the real part,

$$\therefore C_1 = \frac{R_4C_2}{R_3}$$

Comparing the imaginary part,

$$wC_1r_1R_4C_2 = wC_4R_3R_4C_1$$

$$r_1 = \frac{C_4R_3}{C_2}$$

Dissipation factor of capacitor,

$$D = wC_1 r_1 = w \times \frac{R_4 C_2}{R_3} \times \frac{C_4 R_3}{C_2}$$

$$\therefore D = wC_4 R_4$$

Advantages

- ✓ In this type of bridge, the value of capacitance can be measured accurately.
- ✓ It can measure capacitance value over a wide range.
- ✓ It can measure dissipation factor accurately.

Disadvantages

- ✓ It requires two capacitors.
- ✓ Variable standard capacitor is costly.

2.5 Measurements of frequency

2.5.1 Wein's bridge

Wein's bridge is popularly used for measurements of frequency of frequency. In this bridge, the value of all parameters are known. The source whose frequency has to measure is connected as shown in the figure.

$$Z_1 = r_1 + \frac{1}{jwC_1} = \frac{jwC_1 r_1 + 1}{jwC_1}$$

$$Z_2 = \frac{R_2}{1 + jwC_2 R_2}$$

At balance condition, $\dot{Z}_1 \dot{Z}_4 = \dot{Z}_2 \dot{Z}_3$

$$\frac{jwC_1 r_1 + 1}{jwC_1} \times R_4 = \frac{R_2}{1 + jwC_2 R_2} \times R_3$$

$$(1 + jwC_1 r_1)(1 + jwC_2 R_2) R_4 = R_2 R_3 \times jwC_1$$

$$\left[1 + jwC_2 R_2 + jwC_1 r_1 - w^2 C_1 C_2 r_1 R_2\right] = jwC_1 \frac{R_2 R_3}{R_4}$$

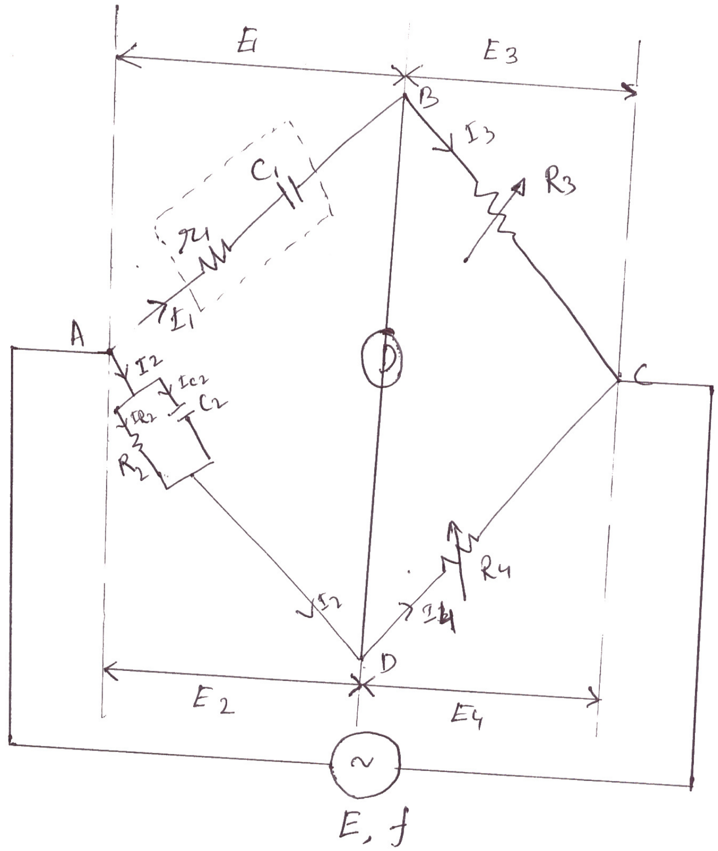


Fig 2.23 Wein's bridge

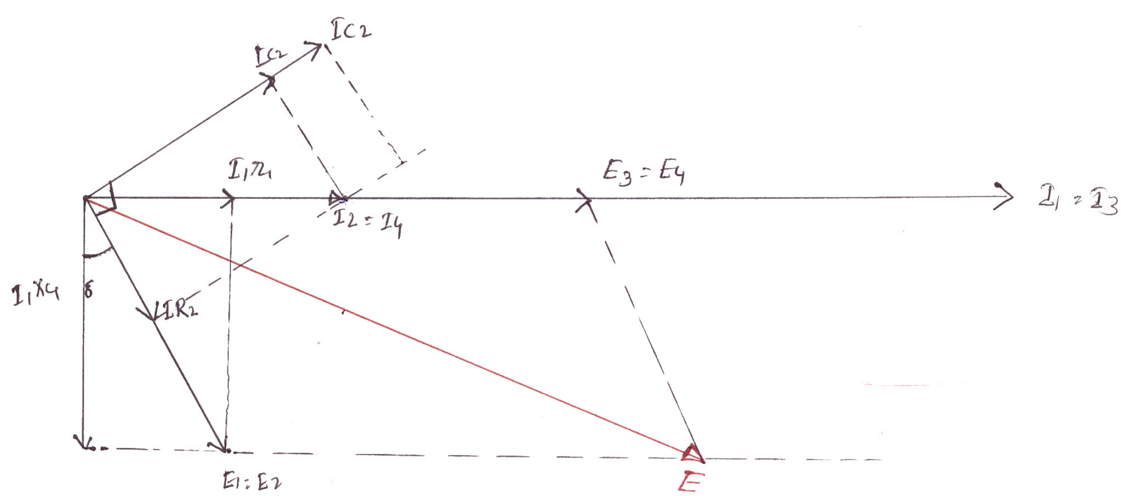


Fig 2.24 Phasor diagram of Wein's bridge

Comparing real term,

$$1-w^2C_1C_2r_1R_2=0$$

$$w^2C_1C_2r_1R_2=1$$

$$w^2=\frac{1}{C_1C_2r_1R_2}$$

$$w=\frac{1}{\sqrt{C_1C_2r_1R_2}},\ f=\frac{1}{2\Pi\sqrt{C_1C_2r_1R_2}}$$

NOTE

The above bridge can be used for measurements of capacitance. In such case, r_1 and C_1 are unknown and frequency is known. By equating real terms, we will get R_1 and C_1. Similarly by equating imaginary term, we will get another equation in terms of r_1 and C_1. It is only used for measurements of Audio frequency.

A.F=20 HZ to 20 KHZ

R.F=>> 20 KHZ

Comparing imaginary term,

$$wC_2R_2+wC_1r_1=wC_1\frac{R_2R_3}{R_4}$$

$$C_2R_2+C_1r_1=\frac{C_1R_2R_3}{R_4} \text{ ....(2.19)}$$

$$C_1=\frac{1}{w^2C_2r_1R_2}$$

Substituting in eqn. (2.19), we have

$$C_2R_2+\frac{r_1}{w^2C_2r_1R_2}=\frac{R_2R_3}{R_4}C_1$$

Multiplying $\frac{R_4}{R_2R_3}$ in both sides, we have

$$C_2R_2\times\frac{R_4}{R_2R_3}+\frac{1}{w^2C_2R_2}\times\frac{R_4}{R_2R_3}=C_1$$

$$C_1 = \frac{C_2 R_4}{R_3} + \frac{R_4}{w^2 C_2 R_2^2 R_3}$$

$$w^2 C_1 r_1 C_2 R_2 = 1$$

$$r_1 = \frac{1}{w^2 C_2 R_2 C_1} = \frac{1}{w^2 C_2 R_2 \left[\frac{C_2 R_4}{R_3} + \frac{R_4}{w^2 C_2 R_2^2 R_3}\right]}$$

$$= \frac{1}{\left[\frac{w^2 {C_2}^2 R_2 R_4}{R_3} + \frac{R_4}{R_2 R_3}\right]}$$

$$\therefore r_1 = \frac{1}{\frac{R_3}{R_4}\left[w^2 C_2^2 R_2 + \frac{1}{R_2}\right]}$$

$$\therefore r_1 = \frac{R_3}{R_4}\left[\frac{1}{(w^2 C_2^2 R_2 + \frac{1}{R_2})}\right]$$

2.5.2 High Voltage Schering Bridge

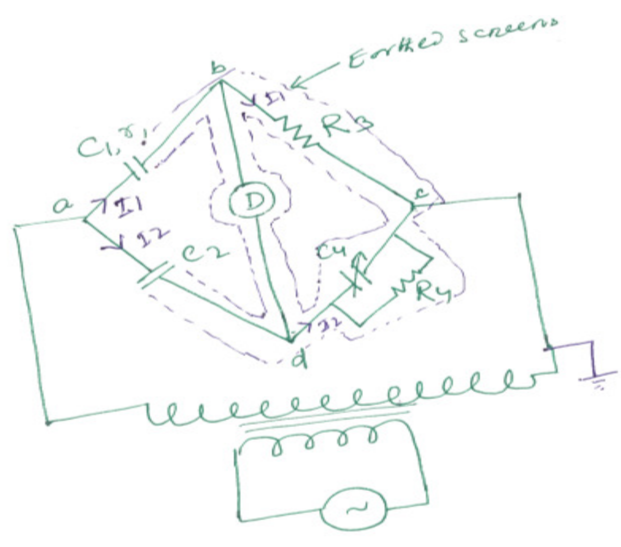


Fig 2.25 High Voltage Schering bridge

(1) The high voltage supply is obtained from a transformer usually at 50 HZ.

2.6 Wagner earthing device:

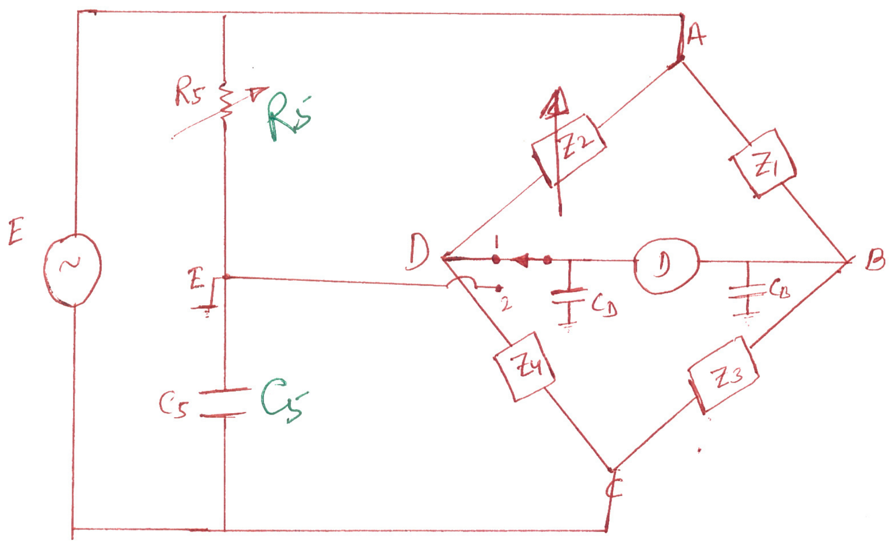


Fig 2.26 Wagner Earthing device

Wagner earthing consists of ‘R’ and ‘C’ in series. The stray capacitance at node ‘B’ and ‘D’ are C_B, C_D respectively. These Stray capacitances produced error in the measurements of ‘L’ and ‘C’. These error will predominant at high frequency. The error due to this capacitance can be eliminated using wagner earthing arm.

Close the change over switch to the position (1) and obtained balanced. Now change the switch to position (2) and obtained balance. This process has to repeat until balance is achieved in both the position. In this condition the potential difference across each capacitor is zero. Current drawn by this is zero. Therefore they do not have any effect on the measurements.

What are the sources of error in the bridge measurements?

- ✓ Error due to stray capacitance and inductance.
- ✓ Due to external field.
- ✓ Leakage error: poor insulation between various parts of bridge can produced this error.
- ✓ Eddy current error.
- ✓ Frequency error.

- ✓ Waveform error (due to harmonics)
- ✓ Residual error: small inductance and small capacitance of the resistor produce this error.

Precaution

- ✓ The load inductance is eliminated by twisting the connecting the connecting lead.
- ✓ In the case of capacitive bridge, the connecting lead are kept apart.($\because C = \frac{A \in 0 \in r}{d}$)
- ✓ In the case of inductive bridge, the various arm are magnetically screen.
- ✓ In the case of capacitive bridge, the various arm are electro statically screen to reduced the stray capacitance between various arm.
- ✓ To avoid the problem of spike, an inter bridge transformer is used in between the source and bridge.
- ✓ The stray capacitance between the ends of detector to the ground, cause difficulty in balancing as well as error in measurements. To avoid this problem, we use wagner earthing device.

2.7 Ballastic galvanometer

This is a sophisticated instrument. This works on the principle of PMMC meter. The only difference is the type of suspension is used for this meter. Lamp and glass scale method is used to obtain the deflection. A small mirror is attached to the moving system. Phosphorous bronze wire is used for suspension.

When the D.C. voltage is applied to the terminals of moving coil, current flows through it. When a current carrying coil kept in the magnetic field, produced by permanent magnet, it experiences a force. The coil deflects and mirror deflects. The light spot on the glass scale also move. This deflection is proportional to the current through the coil.

$$i = \frac{Q}{t}, Q = it = \int idt$$

$\theta \propto Q$, deflection $\propto$ Charge

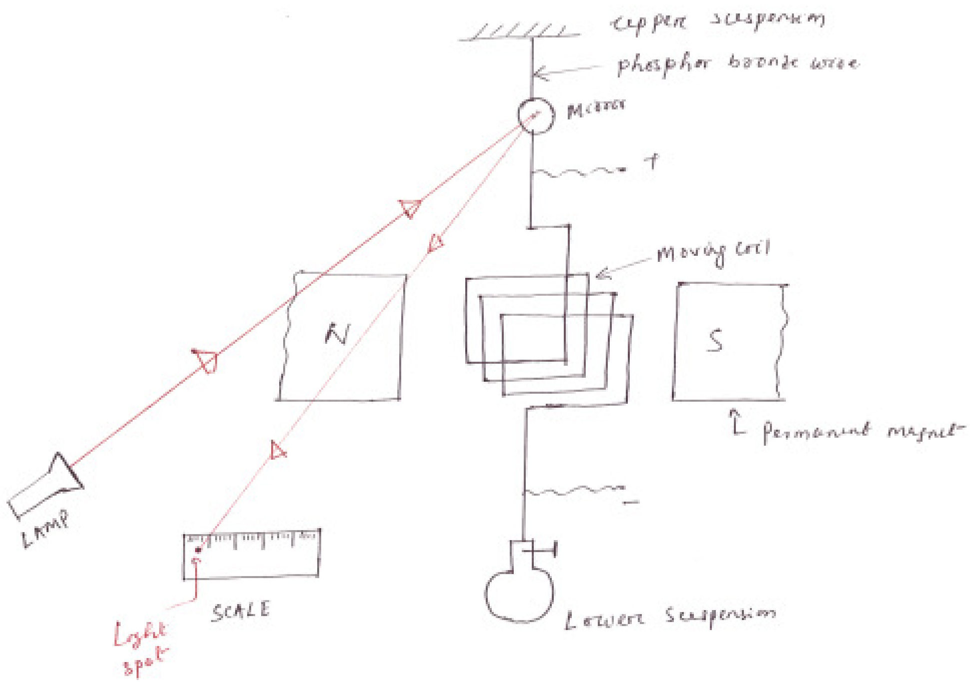


Fig 2.27 Ballastic galvanometer

2.8 Measurements of flux and flux density (Method of reversal)

D.C. voltage is applied to the electromagnet through a variable resistance R_1 and a reversing switch. The voltage applied to the toroid can be reversed by changing the switch from position 2 to position '1'. Let the switch be in position '2' initially. A constant current flows through the toroid and a constant flux is established in the core of the magnet.

A search coil of few turns is provided on the toroid. The B.G. is connected to the search coil through a current limiting resistance. When it is required to measure the flux, the switch is changed from position '2' to position '1'. Hence the flux reduced to zero and it starts increasing in the reverse direction. The flux goes from $+\phi$ to $-\phi$, in time 't' second. An emf is induced in the search coil, science the flux changes with time. This emf circulates a current through R_2 and B.G. The meter deflects. The switch is normally closed. It is opened when it is required to take the reading.

2.8.1 Plotting the BH curve

The curve drawn with the current on the X-axis and the flux on the Y-axis, is called magnetization characteristics. The shape of B-H curve is similar to shape of magnetization characteristics. The residual magnetism present in the specimen can be removed as follows.

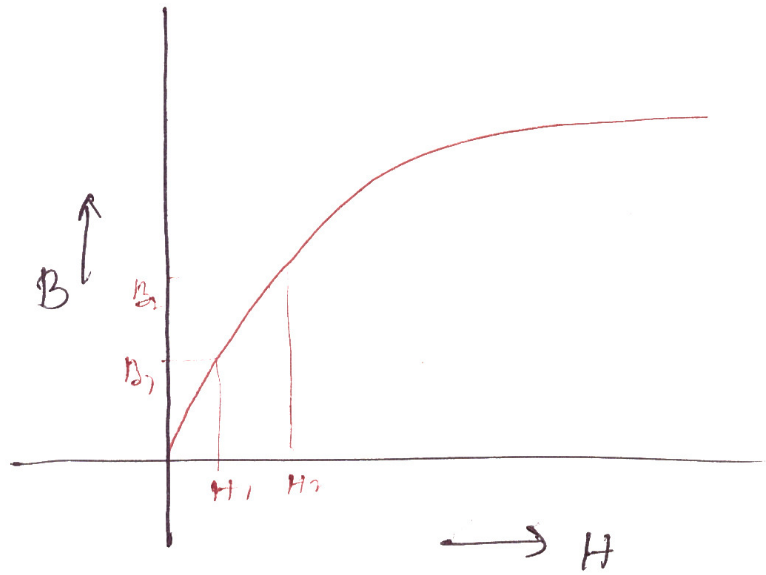


Fig 2.28 BH curve

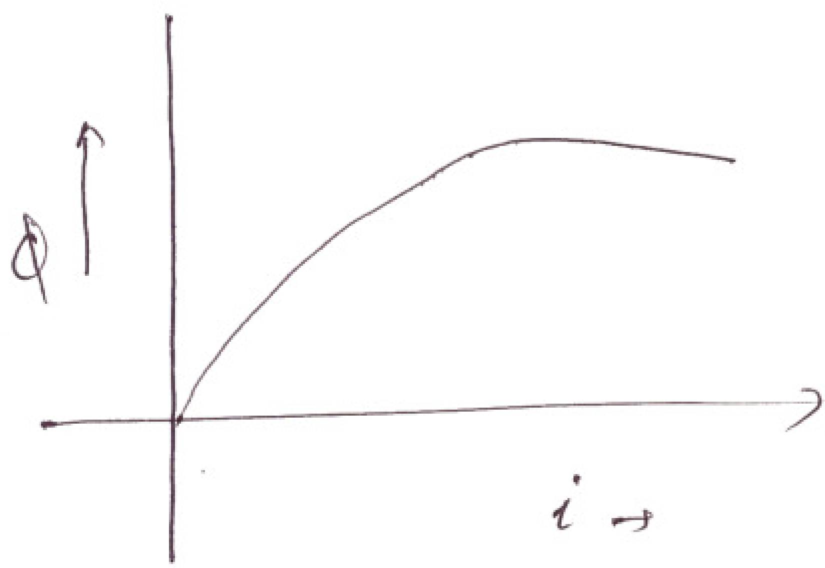


Fig 2.29 Magnetization characteristics

Close the switch ‘S_2’ to protect the galvanometer, from high current. Change the switch S1 from position ‘1’ to ‘2’ and vice versa for several times.

To start with the resistance ‘R_1’ is kept at maximum resistance position. For a particular value of current, the deflection of B.G. is noted. This process is repeated for various value of current. For each deflection flux can be calculated.($B = \frac{\phi}{A}$)

Magnetic field intensity value for various current can be calculated.().The B-H curve can be plotted by using the value of ‘B’ and ‘H’.

2.8.2 Measurements of iron loss:

Let R_P= pressure coil resistance

R_S = resistance of coil S1

E= voltage reading= Voltage induced in S_2

I= current in the pressure coil

V_P= Voltage applied to wattmeter pressure coil.

W= reading of wattmeter corresponding voltage V

W_1= reading of wattmeter corresponding voltage E

$$W \to V \qquad \frac{W_1}{W} = \frac{E}{V} \Rightarrow W_1 = \frac{E \times W}{V}$$
$$W_1 \to E_P$$

W1=Total loss=Iron loss+ Cupper loss.

The above circuit is similar to no load test of transformer.

In the case of no load test the reading of wattmeter is approximately equal to iron loss. Iron loss depends on the emf induced in the winding. Science emf is directly proportional to flux. The voltage applied to the pressure coil is V. The corresponding of wattmeter is ‘W’. The iron loss corresponding E is $E = \frac{WE}{V}$. The reading of the wattmeter includes the losses in the pressure coil and copper loss of the winding S1. These loses have to be subtracted to get the actual iron loss.

2.9 Galvanometers

D-Arsonval Galvanometer

Vibration Galvanometer

Ballistic C

2.9.1 D-arsonval galvanometer (d.c. galvanometer)

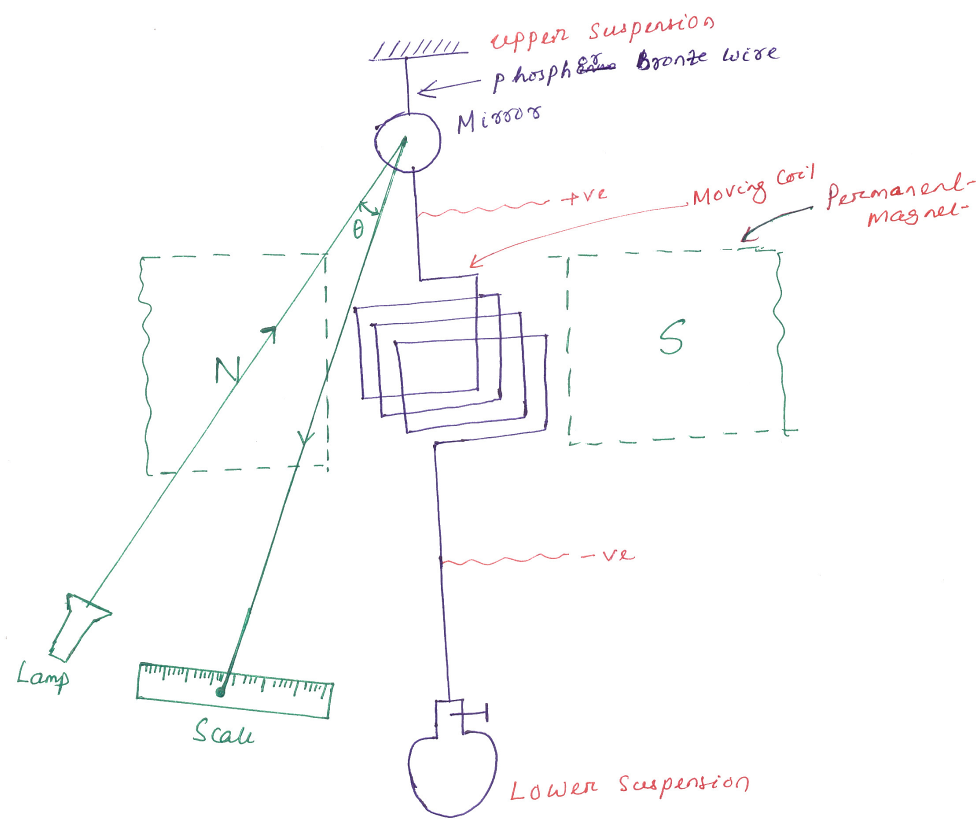


Fig 2.30 D-Arsonval Galvanometer

Galvanometer is a special type of ammeter used for measuring μ A or mA. This is a sophisticated instruments. This works on the principle of PMMC meter. The only difference is the type of suspension used for this meter. It uses a sophisticated suspension called taut suspension, so that moving system has negligible weight.

Lamp and glass scale method is used to obtain the deflection. A small mirror is attached to the moving system. Phosphors bronze is used for suspension.

When D.C. voltage is applied to the terminal of moving coil, current flows through it. When current carrying coil is kept in the magnetic field produced by P.M. , it experiences a force. The light spot on the glass scale also move. This deflection is proportional to the current through the coil. This instrument can be used only with D.C. like PMMC meter.

The deflecting Torque,

T_D=BINA

T_D=GI,   Where G=BAN

$T_C=K_S\theta=S\theta$

At balance, $T_C=T_D \Rightarrow S\theta=GI$

$$\therefore \theta=\frac{GI}{S}$$

Where G= Displacements constant of Galvanometer

S=Spring constant

2.9.2 Vibration Galvanometer (A.C. Galvanometer)

The construction of this galvanometer is similar to the PMMC instrument except for the moving system. The moving coil is suspended using two ivory bridge pieces. The tension of the system can be varied by rotating the screw provided at the top suspension. The natural frequency can be varied by varying the tension wire of the screw or varying the distance between ivory bridge piece.

When A.C. current is passed through coil an alternating torque or vibration is produced. This vibration is maximum if the natural frequency of moving system coincide with supply frequency. Vibration is maximum, science resonance takes place. When the coil is vibrating , the mirror oscillates and the dot moves back and front. This appears as a line on the glass scale. Vibration galvanometer is used for null deflection of a dot appears on the scale. If the bridge is unbalanced, a line appears on the scale

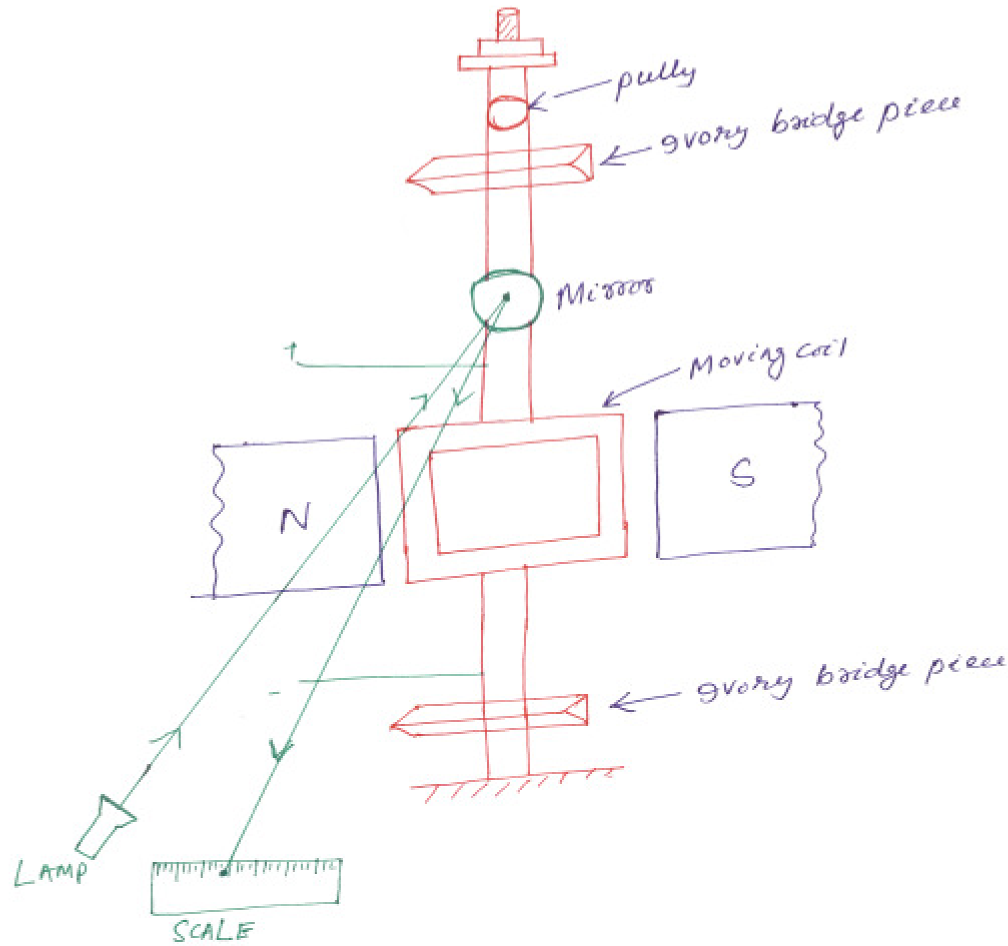


Fig 2.31 Vibration Galvanometer

Example 2.2-In a low- Voltage Schering bridge designed for the measurement of permittivity, the branch 'ab' consists of two electrodes between which the specimen under test may be inserted, arm 'bc' is a non-reactive resistor R_3 in parallel with a standard capacitor C_3, arm CD is a non-reactive resistor R_4 in parallel with a standard capacitor C_4, arm 'da' is a standard air capacitor of capacitance C_2. Without the specimen between the electrode, balance is obtained with following values , C_3=C_4=120 pF, C_2=150 pF, R_3=R_4=5000Ω.With the specimen inserted, these values become C_3=200 pF,C_4=1000 pF,C_2=900 pF and R_3=R_4=5000Ω. In such test w=5000 rad/sec. Find the relative permittivity of the specimen?

Sol: $$\text{Relative permittivity}(\varepsilon_r) = \frac{\text{capacitance measured with given medium}}{\text{capacitance measured with air medium}}$$

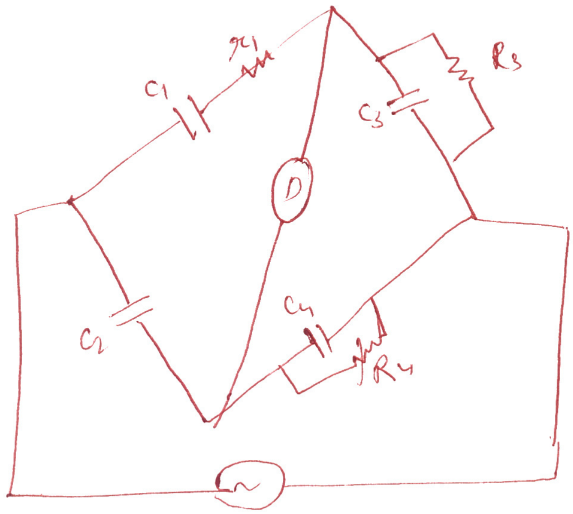


Fig 2.32 Schering bridge

$$C_1 = C_2(\frac{R_4}{R_3})$$

Let capacitance value C_0, when without specimen dielectric.

Let the capacitance value C_S when with the specimen dielectric.

$$C_0 = C_2(\frac{R_4}{R_3}) = 150 \times \frac{5000}{5000} = 150\,pF$$

$$C_S = C_2(\frac{R_4}{R_3}) = 900 \times \frac{5000}{5000} = 900\,pF$$

$$\varepsilon_r = \frac{C_S}{C_0} = \frac{900}{150} = 6$$

Example 2.3- A specimen of iron stamping weighting 10 kg and having a area of 16.8 cm^2 is tested by an episten square. Each of the two winding S_1 and S_2 have 515 turns. A.C. voltage of 50 HZ frequency is given to the primary. The current in the primary is 0.35 A. A voltmeter connected to S_2 indicates 250 V. Resistance of S_1 and S_2 each equal to 40 Ω. Resistance of pressure coil is 80 kΩ. Calculate maximum flux density in the specimen and iron loss/kg if the wattmeter indicates 80 watt?

Sol^n- $E = 4.44 f\phi_m N$

$$B_m = \frac{E}{4.44 fAN} = 1.3 wb/m^2$$

Iron loss= $W(1+\frac{R_S}{R_P}) - \frac{E^2}{(R_S + R_P)}$

$$= 80(1+\frac{40}{80\times10^3}) - \frac{250^2}{(40+80\times10^3)} = 79.26 watt$$

Iron loss/ kg=79.26/10=7.926 w/kg.

8. Resolution

Resolution is the smallest detectable incremental change of input parameter that can be detected in the output signal. Resolution can be expressed either as a proportion of the full-scale reading or in absolute terms. For example, if a LVDT sensor measures a displacement up to 20 mm and it provides an output as a number between 1 and 100 then the resolution of the sensor device is 0.2 mm.

9. Stability

Stability is the ability of a sensor device to give same output when used to measure a constant input over a period of time. The term 'drift' is used to indicate the change in output that occurs over a period of time. It is expressed as the percentage of full range output.

10.Dead band/time

The dead band or dead space of a transducer is the range of input values for which there is no output. The dead time of a sensor device is the time duration from the application of an input until the output begins to respond or change.

11.Repeatability

It specifies the ability of a sensor to give same output for repeated applications of same input value. It is usually expressed as a percentage of the full range output:

Repeatability = (maximum – minimum values given) X 100 / full range (2.1.2)

12.Response time

Response time describes the speed of change in the output on a step-wise change of the measurand. It is always specified with an indication of input step and the output range for which the response time is defined.

Classification of sensors

Sensors can be classified into various groups according to the factors such as measurand, application fields, conversion principle, energy domain of the measurand and thermodynamic considerations. These general classifications of sensors are well described in the references [2, 3].

Detail classification of sensors in view of their applications in manufacturing is as follows.

A. Displacement, position and proximity sensors
    - Potentiometer
    - Strain-gauged element
    - Capacitive element
    - Differential transformers
    - Eddy current proximity sensors
    - Inductive proximity switch
    - Optical encoders
    - Pneumatic sensors
    - Proximity switches (magnetic)
    - Hall effect sensors
B. Velocity and motion
    - Incremental encoder
    - Tachogenerator
    - Pyroelectric sensors

C. Force
    - Strain gauge load cell
D. Fluid pressure
    - Diaphragm pressure gauge
    - Capsules, bellows, pressure tubes
    - Piezoelectric sensors
    - Tactile sensor
    -
E. Liquid flow
    - Orifice plate
    - Turbine meter
F. Liquid level
    - Floats
    - Differential pressure
G. Temperature
    - Bimetallic strips
    - Resistance temperature detectors
    - Thermistors
    - Thermo-diodes and transistors
    - Thermocouples
    - Light sensors
    - Photo diodes
    - Photo resistors

Displacement and position sensors

Displacement sensors are basically used for the measurement of movement of an object. Position sensors are employed to determine the position of an object in relation to some reference point.

Proximity sensors are a type of position sensor and are used to trace when an object has moved with in particular critical distance of a transducer.

Displacement sensors

1. Potentiometer Sensors

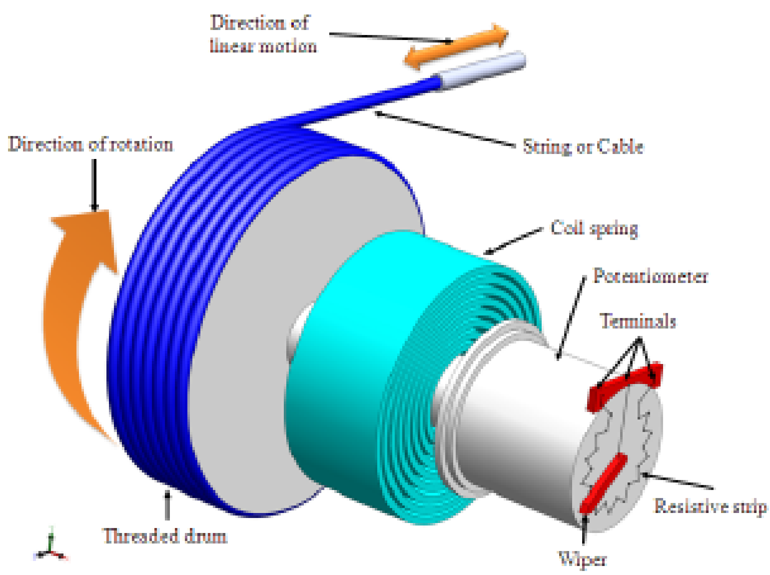


Figure 2.2.1 Schematic of a potentiometer sensor for measurement of linear displacement

Figure 2.2.1 shows the construction of a rotary type potentiometer sensor employed to measure the linear displacement. The potentiometer can be of linear or angular type. It works on the principle of conversion of mechanical displacement into an electrical signal. The sensor has a resistive element and a sliding contact (wiper). The slider moves along this conductive body, acting as a movable electric contact.

The object of whose displacement is to be measured is connected to the slider by using

- a rotating shaft (for angular displacement)
- a moving rod (for linear displacement)
- a cable that is kept stretched during operation

The resistive element is a wire wound track or conductive plastic. The track comprises of large number of closely packed turns of a resistive wire. Conductive plastic is made up of plastic resin embedded with the carbon powder. Wire wound track has a resolution of the order of ± 0.01 % while the conductive plastic may have the resolution of about 0.1 µm.

During the sensing operation, a voltage V_s is applied across the resistive element. A voltage divider circuit is formed when slider comes into contact with the wire. The output voltage (V_A) is measured as shown in the figure 2.2.2. The output voltage is proportional to the displacement of the slider over the wire. Then the output parameter displacement is calibrated against the output voltage V_A.

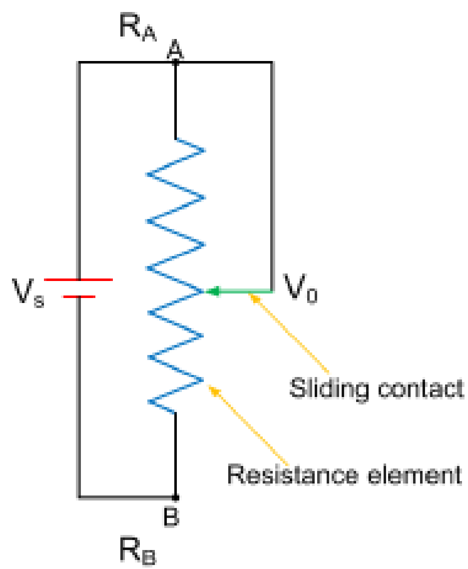


Figure 2.2.2 Potentiometer: electric circuit

$$V_A = I\,R_A \tag{2.2.1}$$

But $I = V_S / (R_A + R_B)$ (2.2.2)

Therefore $V_A = V_S R_A / (R_A + R_B)$ (2.2.3)

As we know that $R = \rho L / A$, where ρ is electrical resistivity, L is length of resistor and A is area of cross section

$$V_A = V_S L_A / (L_A + L_B) \tag{2.2.4}$$

Applications of potentiometer

These sensors are primarily used in the control systems with a feedback loop to ensure that the moving member or component reaches its commanded position.

These are typically used on machine-tool controls, elevators, liquid-level assemblies, forklift trucks, automobile throttle controls. In manufacturing, these are used in control of injection molding machines, woodworking machinery, printing, spraying, robotics, etc. These are also used in computer-controlled monitoring of sports equipment.

2. *Strain Gauges*

The strain in an element is a ratio of change in length in the direction of applied load to the original length of an element. The strain changes the resistance *R* of the element. Therefore, we can say,

$\Delta R/R \propto \varepsilon$;

$$\Delta R/R = G\,\varepsilon \qquad (2.2.5)$$

where G is the constant of proportionality and is called as gauge factor. In general, the value of G is considered in between 2 to 4 and the resistances are taken of the order of 100 Ω.

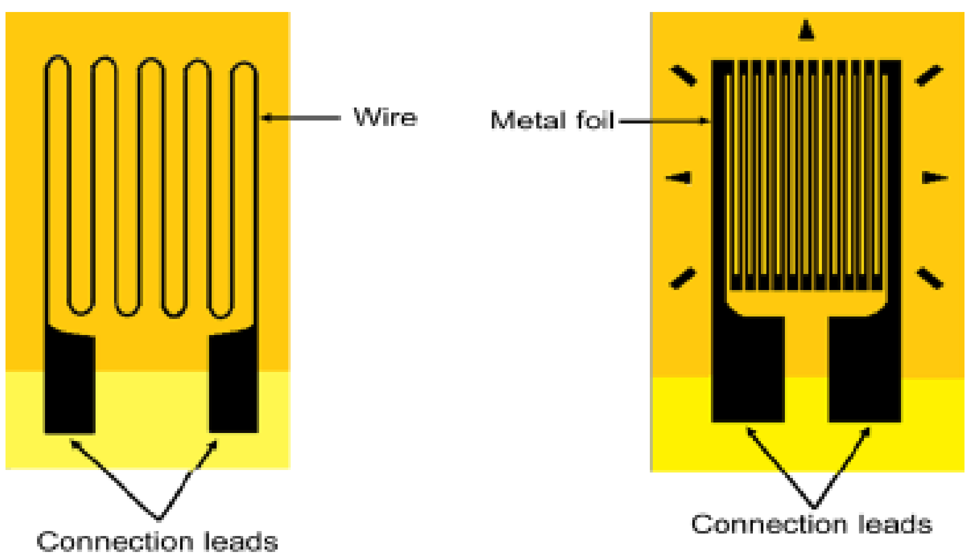


Figure 2.2.3 A pattern of resistive foils

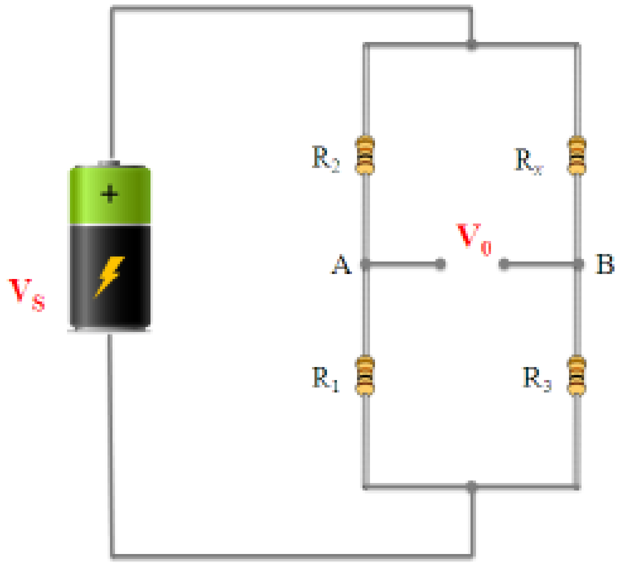


Figure 2.2.4 Wheatstone's bridge

Resistance strain gauge follows the principle of change in resistance as per the equation 2.2.5. It comprises of a pattern of resistive foil arranged as shown in Figure 2.2.3. These foils are made of Constantan alloy (copper-nickel 55-45% alloy) and are bonded to a backing material plastic (ployimide), epoxy or glass fiber reinforced epoxy. The strain gauges are secured to the workpiece by using epoxy or Cyanoacrylate cement Eastman 910 SL. As the workpiece undergoes change in its shape due to external loading, the resistance of strain gauge element changes. This change in resistance can be detected by a using a Wheatstone's resistance bridge as shown in Figure 2.2.4. In the balanced bridge we can have a relation,

$$R_2 / R_1 = R_x / R_3 \qquad (2.2.6)$$

where R_x is resistance of strain gauge element, R_2 is balancing/adjustable resistor, R_1 and R_3 are known constant value resistors. The measured deformation or displacement by the stain gauge is calibrated against change in resistance of adjustable resistor R_2 which makes the voltage across nodes A and B equal to zero.

Applications of strain gauges

Strain gauges are widely used in experimental stress analysis and diagnosis on machines and failure analysis. They are basically used for multi-axial stress fatigue testing, proof testing, residual stress and vibration measurement, torque measurement, bending and deflection measurement, compression and tension measurement and strain measurement.

Strain gauges are primarily used as sensors for machine tools and safety in automotives. In particular, they are employed for force measurement in machine tools, hydraulic or pneumatic press and as impact sensors in aerospace vehicles.

3. *Capacitive element based sensor*

Capacitive sensor is of non-contact type sensor and is primarily used to measure the linear displacements from few millimeters to hundreds of millimeters. It comprises of three plates, with the upper pair forming one capacitor and the lower pair another. The linear displacement might take in two forms:

a. one of the plates is moved by the displacement so that the plate separation changes
b. area of overlap changes due to the displacement.

Figure 2.2.5 shows the schematic of three-plate capacitive element sensor and displacement measurement of a mechanical element connected to the plate 2.

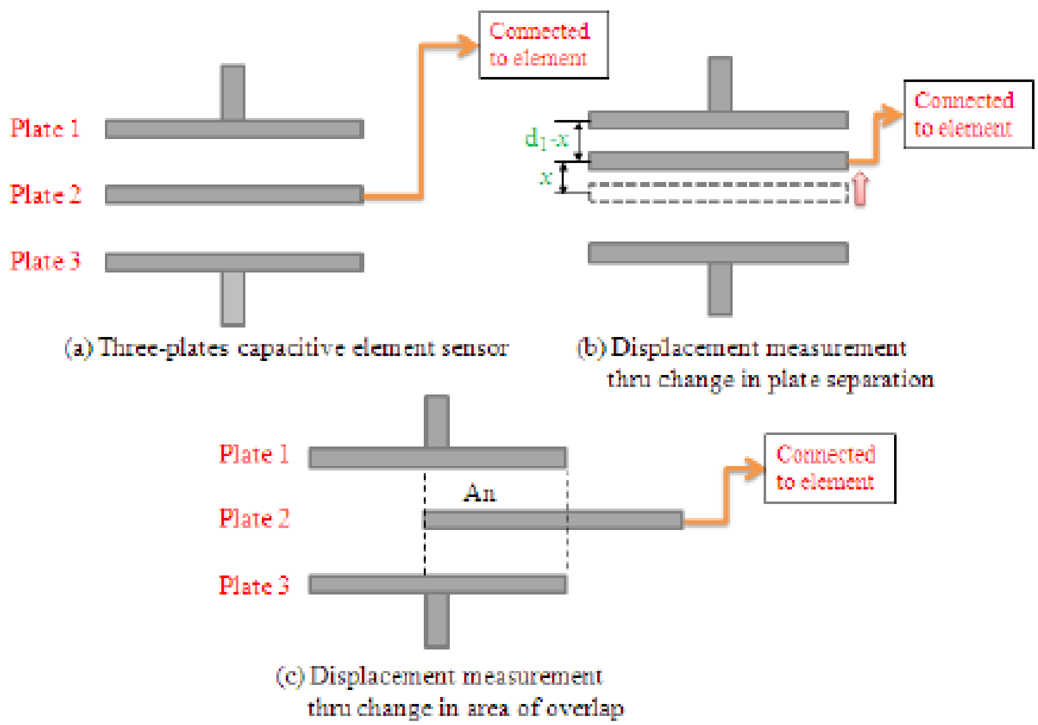


Figure 2.2.5 Displacement measurement using capacitive element sensor

The capacitance C of a parallel plate capacitor is given by,

$$C = \varepsilon_r \varepsilon_o A / d \tag{2.2.7}$$

where ε_r is the relative permittivity of the dielectric between the plates, ε_o permittivity of free space, A area of overlap between two plates and d the plate separation.

As the central plate moves near to top plate or bottom one due to the movement of the element/workpiece of which displacement is to be measured, separation in between the plate changes. This can be given as,

$$C_1 = (\varepsilon_r \, \varepsilon_o \, A) / (d + x) \qquad (2.2.8)$$

$$C_2 = (\varepsilon_r \, \varepsilon_o \, A) / (d - x) \qquad (2.2.9)$$

When C1 and C2 are connected to a Wheatsone's bridge, then the resulting out-of-balance voltage would be in proportional to displacement x.

Capacitive elements can also be used as proximity sensor. The approach of the object towards the sensor plate is used for induction of change in plate separation. This changes the capacitance which is used to detect the object.

Applications of capacitive element sensors

- Feed hopper level monitoring
- Small vessel pump control
- Grease level monitoring
- Level control of liquids
- Metrology applications
  - to measure shape errors in the part being produced
  - to analyze and optimize the rotation of spindles in various machine tools such as surface grinders, lathes, milling machines, and air bearing spindles by measuring errors in the machine tools themselves
- Assembly line testing
  - to test assembled parts for uniformity, thickness or other design features
  - to detect the presence or absence of a certain component, such as glue etc.

4. *Linear variable differential transformer (LVDT)*

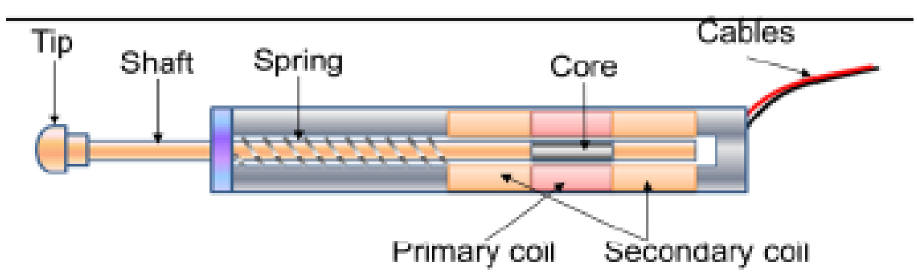


Figure 2.2.6 Construction of a LVDT sensor

Linear variable differential transformer (LVDT) is a primary transducer used for measurement of linear displacement with an input range of about ± 2 to ± 400 mm in general. It has non-linearity error ± 0.25% of full range. Figure 2.2.6 shows the construction of a LVDT sensor. It has three coils symmetrically spaced along an insulated tube. The central coil is primary coil and the other two are secondary coils. Secondary coils are connected in series in such a way that their outputs oppose each other. A magnetic core attached to the element of which displacement is to be monitored is placed inside the insulated tube.

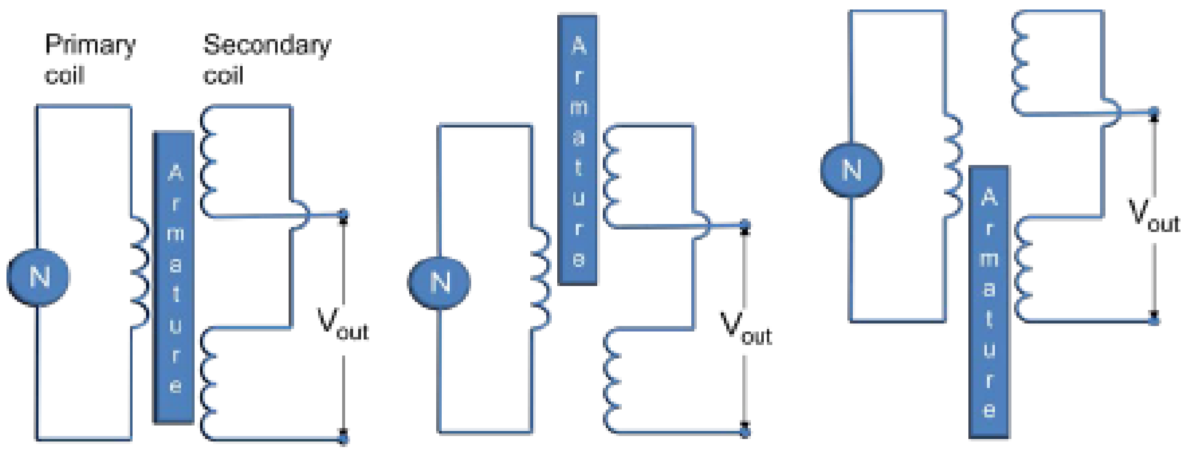


Figure 2.2.7 Working of LVDT sensor

Due to an alternating voltage input to the primary coil, alternating electro-magnetic forces (emfs) are generated in secondary coils. When the magnetic core is centrally placed with its half portion in each of the secondary coil regions then the resultant voltage is zero. If the core is displaced from the central position as shown in Figure 2.2.7, say, more in secondary coil 1 than in coil 2, then more emf is generated in one coil i.e. coil 1 than the other, and there is a resultant voltage from the coils. If the magnetic core is further displaced, then the value of resultant voltage increases in proportion with the displacement. With the help of signal processing devices such as low pass filters and demodulators, precise displacement can be measured by using LVDT sensors.

LVDT exhibits good repeatability and reproducibility. It is generally used as an absolute position sensor. Since there is no contact or sliding between the constituent elements of the sensor, it is highly reliable. These sensors are completely sealed and are widely used in Servomechanisms, automated measurement in machine tools.

A rotary variable differential transformer (RVDT) can be used for the measurement of rotation. Readers are suggested to prepare a report on principle of working and construction of RVDT sensor.

Applications of LVDT sensors

- Measurement of spool position in a wide range of servo valve applications
- To provide displacement feedback for hydraulic cylinders
- To control weight and thickness of medicinal products viz. tablets or pills
- For automatic inspection of final dimensions of products being packed for dispatch
- To measure distance between the approaching metals during Friction welding process
- To continuously monitor fluid level as part of leak detection system
- To detect the number of currency bills dispensed by an ATM

Quiz:

1. Explain the principle of working of LVDT.
2. Describe the working of RVDT with a neat sketch.
3. List the applications of potentiometer sensor in/around your home and office/university.

References

1. Boltan, W., Mechatronics: electronic control systems in mechanical and electrical engineering, Longman, Singapore, 1999.

6. *Hall effect sensor*

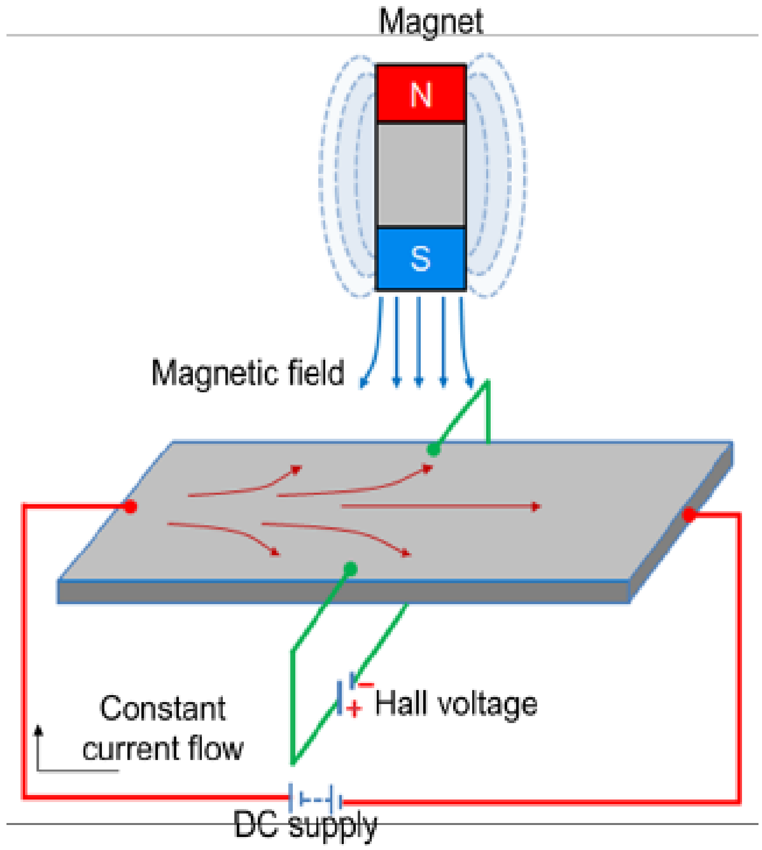


Figure 2.3.8 Principle of working of Hall effect sensor

Figure 2.3.8 shows the principle of working of Hall effect sensor. Hall effect sensors work on the principle that when a beam of charge particles passes through a magnetic field, forces act on the particles and the current beam is deflected from its straight line path. Thus one side of the disc will become negatively charged and the other side will be of positive charge. This charge separation generates a potential difference which is the measure of distance of magnetic field from the disc carrying current.

The typical application of Hall effect sensor is the measurement of fluid level in a container. The container comprises of a float with a permanent magnet attached at its top. An electric circuit with a current carrying disc is mounted in the casing. When the fluid level increases, the magnet will come close to the disc and a potential difference generates. This voltage triggers a switch to stop the fluid to come inside the container.

RTDs are used in the form of thin films, wire wound or coil. They are generally made of metals such as platinum, nickel or nickel-copper alloys. Platinum wire held by a high-temperature glass adhesive in a ceramic tube is used to measure the temperature in a metal furnace. Other applications are:

- Air conditioning and refrigeration servicing
- Food Processing
- Stoves and grills
- Textile production
- Plastics processing
- Petrochemical processing
- Micro electronics
- Air, gas and liquid temperature measurement in pipes and tanks
- Exhaust gas temperature measurement

3. *Thermistors*

Thermistors follow the principle of decrease in resistance with increasing temperature. The material used in thermistor is generally a semiconductor material such as a sintered metal oxide (mixtures of metal oxides, chromium, cobalt, iron, manganese and nickel) or doped polycrystalline ceramic containing barium titanate (BaTiO3) and other compounds. As the temperature of semiconductor material increases the number of electrons able to move about increases which results in more current in the material and reduced resistance. Thermistors are rugged and small in dimensions. They exhibit nonlinear response characteristics.

Thermistors are available in the form of a bead (pressed disc), probe or chip. Figure 2.5.4 shows the construction of a bead type thermistor. It has a small bead of dimension from 0.5 mm to 5 mm coated with ceramic or glass material. The bead is connected to an electric circuit through two leads. To protect from the environment, the leads are contained in a stainless steel tube.

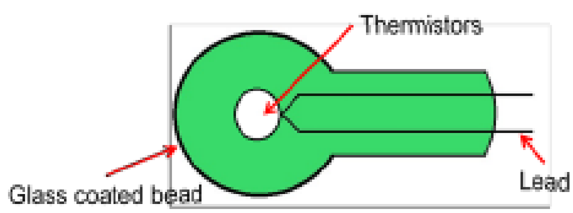


Figure 2.5.4 Schematic of a thermistor

The Oscilloscope: Operation and Applications

1. The Oscilloscope

Oscilloscope Operation (X vs Y mode)

An oscilloscope can be used to measure voltage. It does this by measuring the voltage drop across a resistor and in the process draws a small current. The voltage drop is amplified and used to deflect an electron beam in either the X (horizontal) or Y (vertical) axis using an electric field. The electron beam creates a bright dot on the face of the Cathode Ray Tube (CRT) where it hits the phosphorous. The deflection, due to an applied voltage, can be measured with the aid of the calibrated lines on the graticule.

First we will consider the circuitry that amplifies and conditions the voltage to be measured (the "Amp" block in figure 1).

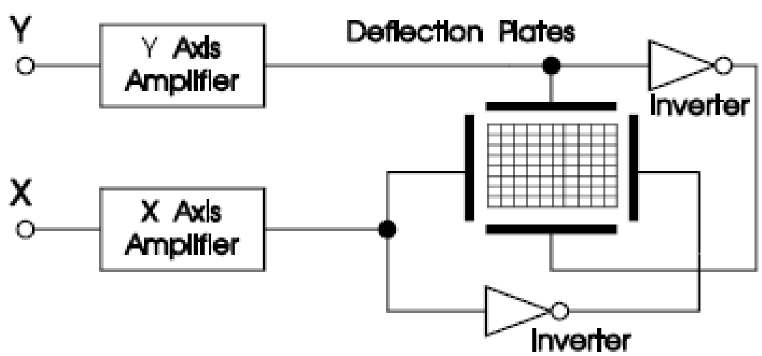


Figure 1. X vs. Y Deflection Block Diagram of the CRT

The deflection of the oscilloscope beam is proportional to the input voltage (after ac or dc coupling). The amount of deflection (Volts/Division) depends upon the setting of the AMPL/DIV control for that channel (see figure 2).

The input signal can be ac or dc coupled. Ac coupling involves adding a series capacitor. This has the effect of blocking (removing) the dc bias and low frequency components of a signal. Dc coupling does not have this problem and therefore allows you to measure voltages right down to 0 Hz. Ac coupling is useful when you are trying to measure a small ac voltage that is "on-top" of a large dc voltage. A typical example is trying to measure the noise of a dc power supply.

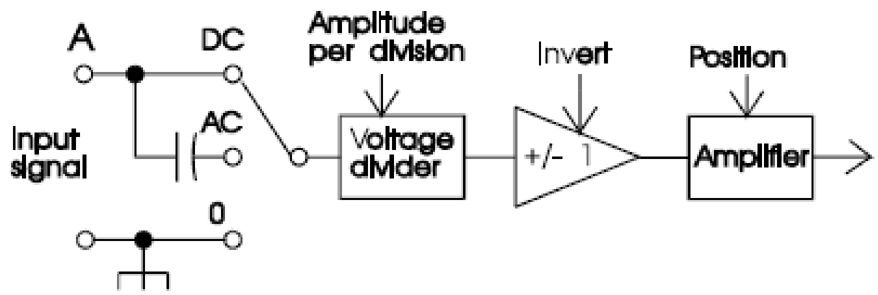


Figure 2. Amplifier Block Diagram

Amplifier Features

AMPL/DIV - This abbreviated name varies but it is generally some short form of amplitude per division. The control is a simple voltage divider (attenuator) which is used to change the sensitivity of the oscilloscope. At a 1 volt/DIV setting, a deflection of one major division on the graticule represents a one volt change at the oscilloscope input.

Calibrated voltage measurements

The small knob within the AMPL/DIV control must be rotated clockwise into its detente position for the amplifiers to be calibrated. Otherwise the voltage/division will be some unknown value greater than what the dial indicates.

INV - There is almost always a control which lets you invert one channel. This can be used along with the ADD function to subtract two voltages. This is necessary because the common input (black lead of the oscilloscope cable) can only be connected to a 0V node. If channel A has V1 + V2 and channel B has voltage V1 then the reading of channel A + (-channel B) = (V1 + V2) + (-V1) = V2

Position - For each axis there is a control which lets you shift the electron beam. With this you can set the zero voltage point to anywhere that is convenient for you.

Oscilloscope Inputs

The input of the oscilloscope can usually be modelled as a resistance and a parallel capacitance (see figure 3). The resistance is usually 1MΩ but it and the capacitance can vary greatly. The total or effective capacitance includes the oscilloscope circuitry (approx. 30 pF), cables (approx. 30 pF/m) and stray capacitance. The resistance will draw current from the circuit while the capacitance will add an RC time constant with its associated time delay, frequency response and distortion of some waveforms.

The common connection (black lead or shield) at the input of the oscilloscope goes to the metal case as the symbol by the input connector shows. Because of this, the common input can only be connected to a 0V point in the circuit. Since the common inputs for both the A and B channels are connected to the case, they are effectively shorted together.

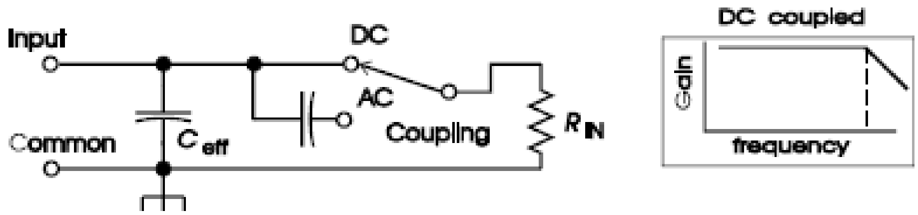


Figure 3. DC-coupled Oscilloscope Input Circuit and Frequency Response

Frequency response is calculated or measured by applying a pure sinusoidal waveform to a circuit. The circuit response is the output voltage divided by the input voltage. This is a complex number that can also be expressed as a magnitude (gain) and phase.

Due to limitations in the amplifiers, the oscilloscope's frequency response is limited. The manufacturer simply lists the half-power point for the oscilloscope without any external effects. Half power is also called the -3dB point. At this point, the voltage has decreased to 70.7% of its maximum. This means that only one-half of the maximum power would be dissipated in a resistive load. Keep in mind that an oscilloscope that is rated at 20 MHz is usually only accurate to 4 MHz for non-sinusoidal waveforms before distortion becomes a problem.

With ac coupling (figure 4), an oscilloscope has another series RC circuit. It acts like a high pass filter (HPF). If you are viewing low frequency signals when ac coupled, not only will you not be able to measure any dc offset, but you will also be removing some low frequency information.

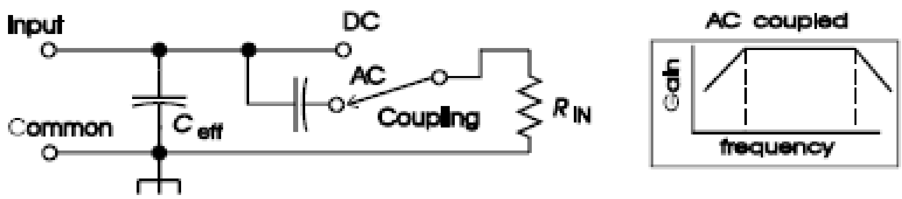


Figure 4. AC-coupled Oscilloscope Input Circuit and Frequency Response

Oscilloscope Operation (Voltage vs Time)

The main function of an oscilloscope is to show voltage vs time. This is done by applying a ramp (or sawtooth) waveform into the X-axis amplifier as shown in figure 5. During the rising edge of the ramp, the electron beam scans across the screen. When the voltage drops back to 0V, the beam is turned off and quickly goes back to its starting point. This is signified by a thick line when the beam is on and a thin one when it is off (blanked).

To obtain a stable picture on the CRT screen, the ramp waveform has to be in phase with the signal that you want to observe. This is done with a triggering circuit. The triggering circuit allows the oscilloscope to draw repeatedly the same waveform over and over by identifying the same point on a repetitive waveform.

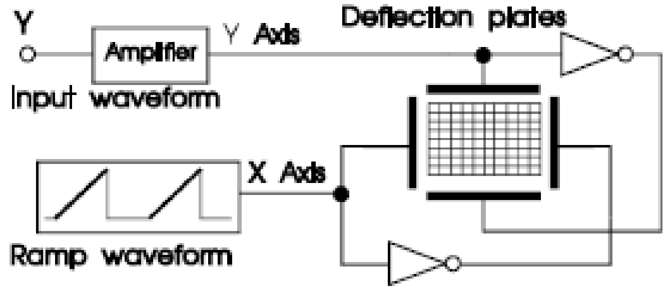


Figure 5. A Ramp-driven X-axis input

The triggering circuit allows you to select a voltage (an analog value) and an edge or slope (positive or negative) for the triggering circuit to compare to the input waveform. When the two are equal, the circuit puts out a pulse. This pulse triggers the ramp waveform generator to do one cycle of its rising and falling edges. Once the ramp has started a cycle of increasing voltage, it can not be retriggered until it has completed the full ramp and returned to 0V. This is illustrated in figure 6 for a single cycle and in figure 7 for multiple cycles.

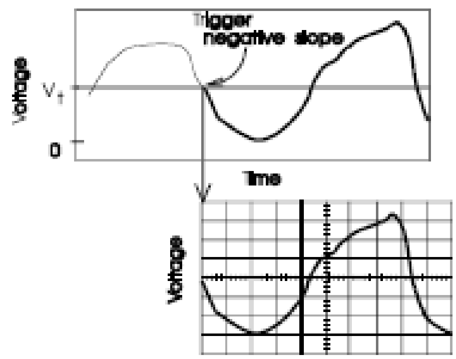


Figure 6. A Triggering Example

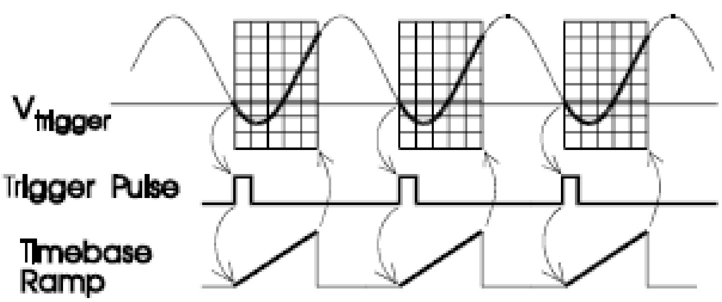


Figure 7. Several Triggering Cycles

Not only do you have control over the starting point of the ramp, but the amount of time that the ramp takes to reach its maximum voltage (the right hand side of the CRT screen) can be adjusted with the timebase control. In essence, you have a "window". You can move the window to any point on a waveform with the triggering circuit and you can change the size of the window with the timebase.

The time-base control allows you to set the time / division that the beams takes to scan across the screen. Just like the voltage selector, there is a calibration knob in the middle of the control. Unless the vernier (calibration knob) is 'clicked' in to its most clockwise position, the time per division is unknown.

When set to AUTO (automatic) triggering, the oscilloscope will always show a trace. However, when you use a manual triggering mode (DC, AC), many strange things can happen. For example, if the triggering voltage or level is set to +10V and the waveform never exceeds +5V, the triggering circuit will never trigger and the screen will stay blank.

You may think that in a condition of no triggering, you would still have a bright dot on the screen because the electron beam would go to its 'home' or undeflected position. Since the oscilloscope is designed to work with a moving electron beam, a stationary beam can very quickly 'burn' a hole in the phosphorous coating of the screen. To prevent this, there is a ' blanking' circuit which turns off the electron beam. Blanking occurs when there is no triggering or when the electron beam is sweeping from the right edge back to the left side of the screen.

Time measurements are done the same way as voltage measurements. As long as the timebase is calibrated you multiply the number of divisions by the number of seconds per division to get the total time difference. Phase measurements are done by comparing the measured time to the period of the waveform.

Oscilloscope Two Channel Operation
You can view two voltage waveforms at once by using two Y-axis (vertical) input channels. The individual channels are sometimes labelled as '1' and '2' or as 'A' and 'B'. Since there is only one electron beam, you have to share its drawing time between both waveforms. This may be accomplished using either the **chop** or **alternate** modes.

When in the chop mode (figure 8), the oscilloscope displays a little bit of channel A, then a little bit of B, then A, then B ....during a single sweep of the electron beam. If you increase the timebase to about 1µs/division, you can start to see the individual pieces as it chops between one channel and the other channel.

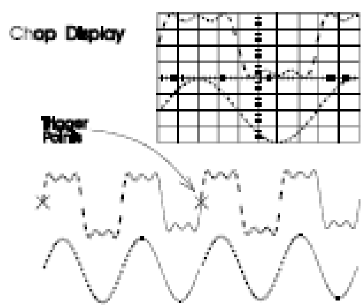


Figure 8. Chop Mode

In the alternate mode (figure 9), the oscilloscope will sweep the electron beam twice across the screen. The first time it will draw the signal from channel A and the next time from channel B. At very low timebase settings, you can see it draw one channel and then the other in successive passes.
Note: When you use the alternate function, the two waveforms that you see are from different points in time and the triggering circuit has to trigger twice.

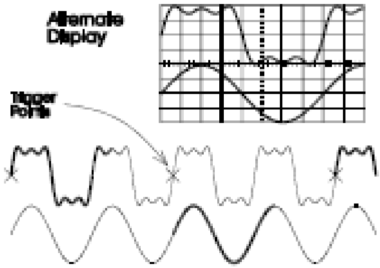


Figure 9. Alternate Mode

The reason that you can see a non-flickering image on the screen is because the phosphorous coating on the CRT has persistence. In essence, the phosphorous acts like a low pass filter and averages several images that are drawn on the screen.

By viewing two signals at a time, you can measure relative time differences. By combining a voltage and phase measurement (relative to the appropriate reference), you can measure a phasor value.

With a two channel oscilloscope, you have the ability to trigger on each waveform and electronically switch (chop or alt) between them as well. A block diagram of a oscilloscope has now become as shown in figure 10.

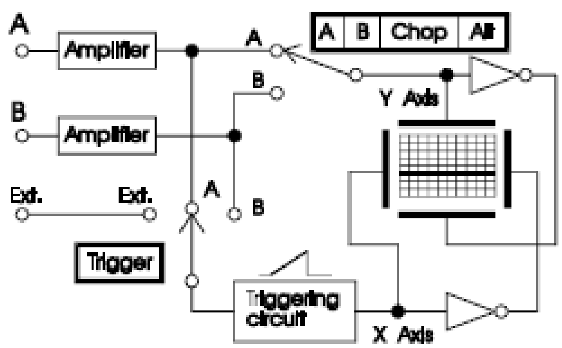


Figure 10. A Simple Oscilloscope Block Diagram

Some oscilloscopes offer a way to alternately trigger as depicted in figure 11. When combined with alternate displaying, you can stably display two waveforms of any frequency by alternately showing each channel and triggering on the channel that is being drawn. This way, the oscilloscope is acting like a two beam scope with both waveforms triggering at the same voltage and slope. However, there is no way to know what the relationship is between one waveform and the other when using alternate triggering.

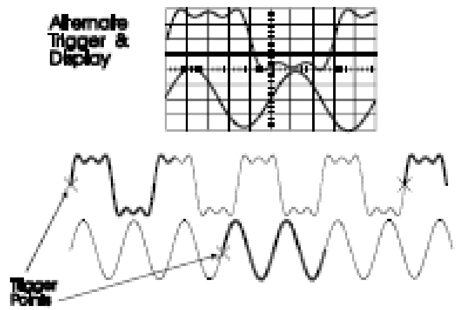


Figure 11. Stable Triggering of Two Different Frequencies

If you have two waveforms that do not have the same frequency, it is still possible to show them as two stable waveforms on a normal oscilloscope. In figure 11, you will notice that if the triggering occurs at the 'X', both waveforms are in phase (ie. at the same phase each time the timebase triggers). The condition for a stable display is not that two waveforms have to be of exactly the same frequency, but that when they are triggered, they have to be in phase.
Or $n \cdot f1 = m \cdot f2$ where n, m are integers. That is not necessary, but it is sufficient. There are many other ways to achieve a stable trace when you consider that the trigger circuit will wait for the next triggering point. There is also a control on some oscilloscopes, called 'hold off', which allows you to add a delay between the end of the trace being drawn and the time when the triggering circuit starts to look for the next triggering point. That can be used to stabilize the display under some circumstances.

Remember that all of this applies only for repetitive waveforms that are properly triggered. If the triggering is not stable, or the waveform is not repetitive, you will see a constantly moving image or several images offset and superimposed.

A slightly more complicated block diagram of an oscilloscope, with the typical functions found in the laboratory, is illustrated in figure 12.

Accuracy
There are many factors affecting the accuracy of oscilloscope measurements.
There are errors due to the input channel voltage divider, timebase control, the use of magnifiers, the accuracy to which the CRT deflection can be read, beam thickness, temperature etc. The voltage divider error will be the same for all readings that are done on the same timebase and voltage range, but may be different each time the range is changed. Measurements over only two divisions can incur two to three times the error of those made over the centre eight divisions.

If the phase angle is used in a trigonometric function, this error can be multiplied by the slope of the function. Consider that the tangent of a 1% phase error entered at 85 degrees is much worse (20%) than the same 1% error on a sine function (0.2%) at the same angle. To get a feel for this look at the Taylor expansion of the trigonometric functions.

It is wise to consult the user manual for a particular instrument's accuracy specifications.

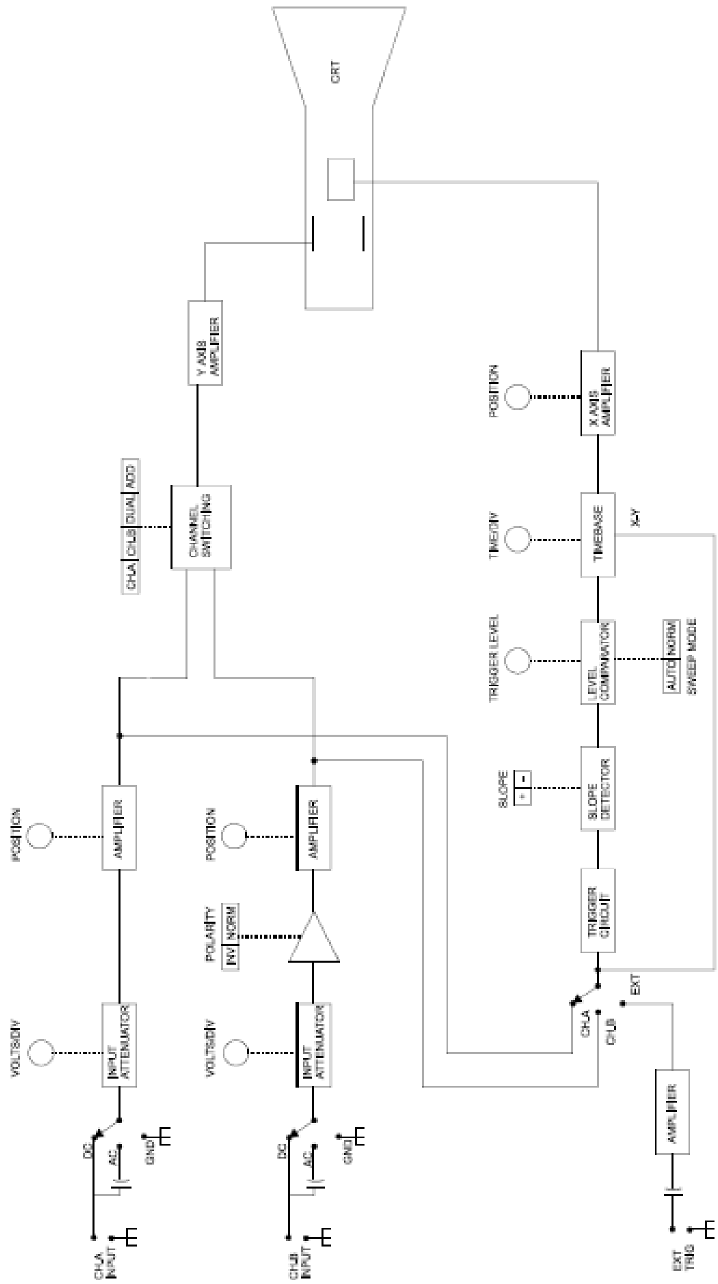


Figure 12. Oscilloscope Block Diagram

2. Iwatsu Model SS-5702 Oscilloscope

Accuracy

The Iwatsu SS-5702 oscilloscope is used in this laboratory. All measurements on the graticule should be made over as many divisions as possible. For simplicity, assume the Iwatsu SS-5702 oscilloscope's error is ±5% for a measurement on either the vertical or horizontal scales over eight divisions.

Front Panel Controls

The front panel controls, shown in figure 13, will be described in the remainder of this section.

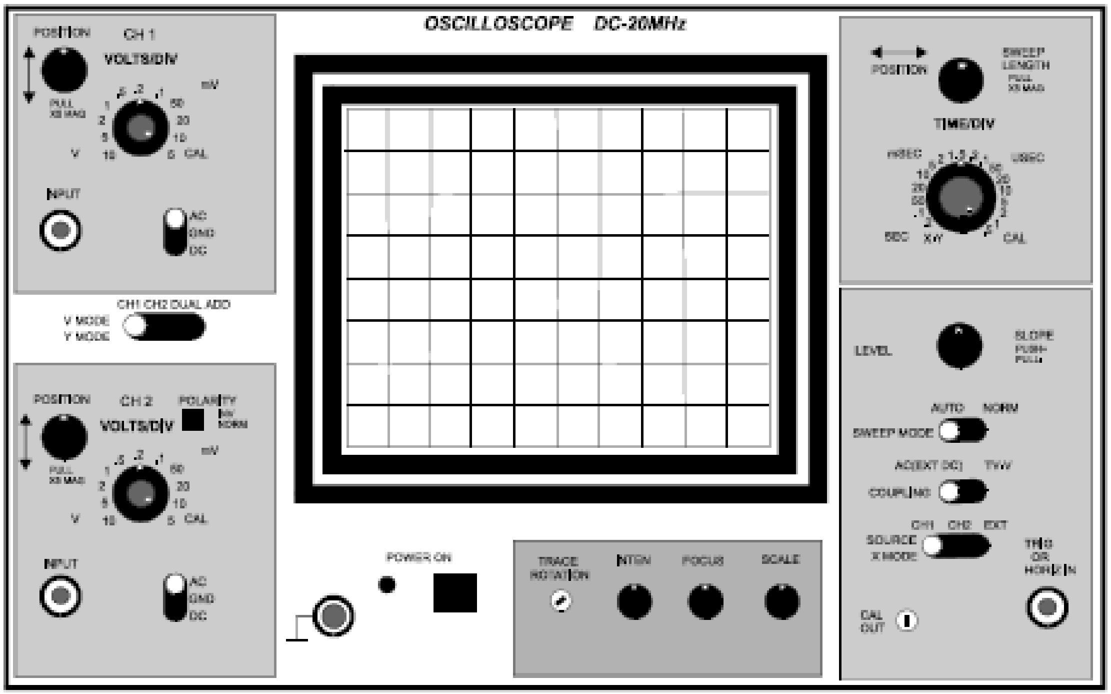


Figure 13. The Iwastu SS-5702 Oscilloscope

Front Panel Controls

The power, trace rotation, intensity, focus, and scale illumination controls are located at the bottom centre.
The vertical controls and input selection are on the left side.
The horizontal controls and horizontal input selection are on the right side.
The triggering selection section is at the bottom right.

The power on/off, trace rotation, intensity, focus and scale illumination

Trace rotation has to be checked with just a straight line across the screen of the oscilloscope.

The intensity should be adjusted to a mid point (or more clockwise).
The focus of the beam can be done while observing the straight line display.
The scale illumination can be adjusted to the operator's preference.

Vertical Inputs
Vertical channels are used for measuring voltage.
The beam is deflected vertically as a result of the signal being applied to the vertical input of the channel (CH).

CH-1 and **CH-2** are the labels for the two vertical inputs.

Each channel has a position control, a range selection switch, a pull "x5" knob, a coupling selector switch (AC/GND/DC), and an input connector.

Each channel Range switch (VOLTS/DIV) has a smaller knob in the middle of the Range Selector Switch. And there is an arrow showing that the Range Selector Switch is in the "**CAL**" position when rotated fully clockwise.

In the centre of the two channel sections is the channel selection switch. You can choose to have CH-1, CH-2, both (DUAL), or ADD.

In addition, **CH-2** has a **"Polarity"** switch. You push the polarity switch in to **"INVERT"** the polarity of the signal being displayed. The **"NORM"** or out position is the normal position of the polarity switch.

Horizontal Inputs
The **"EXT"** (HORIZONTAL IN) can be supplied a voltage directly via the connector at the bottom right of the panel, or the Horizontal can be driven by an internal timebase circuit which generates the voltage.

Timebase
In the timebase mode, the horizontal signal is from an internal source which changes **linearly with respect to time**. Hence, the beam is deflected to give us a calibration of time for the horizontal scale.

The position control, the pull "x5" magnifier switch, the time range selection switch, and the range "CAL" knob all affect the X-axis of the display.

Trigger Sources
The TRIGGER SOURCE may be selected from one of three sources, **CH-1, CH-2, or EXT.** Look at the bottom right of the control panel.

Calibration Source
The Calibration Source is an internal source, available on the oscilloscope.
Look at the bottom right side of the front panel. The output is labelled 0.3 V.
This is a 1000 Hz, square-wave (the 50 % duty cycle is not accurate).

3. Oscilloscope Applications

Voltage and Time Measurements

Note: **The oscilloscope measures divisions of deflection not voltage or time.** From the divisions of deflection you can calculate the time or voltage.

Differential Measurements

An important application of the oscilloscope is differential measurements. Such measurements are necessary because both vertical channels have one terminal connected to the chassis common (ie single ended). To measure a floating (off ground) voltage you have to use the "invert and add" feature of the oscilloscope. For example, in figure 14, to measure V1:

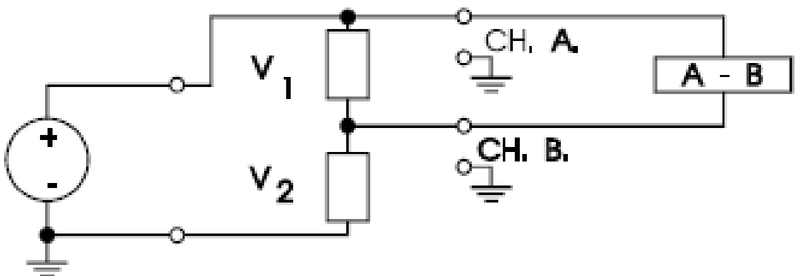


Figure 14. Differential Measurement Example

Channel A measures (V1 + V2) relative to ground while channel B measures V2 relative to ground. By pushing the invert button you negate the voltage displayed on channel B. Then you can add the channels together with the "ADD" display mode. The waveform now displayed is (V1 + V2) + (- V2) = V1.

Bandwidth (-3 dB) Measurement

This measurement is easily done by first finding the maximum gain (max. VOUT / VIN at frequency ω_o) and **adjusting the oscilloscope** so that the sinewave fills seven divisions peak to peak. A -3dB point can be found by increasing and/or decreasing the frequency until the gain is reduced to $\sqrt{2}$. **If the input voltage has remained constant** this will occur when the output voltage is five divisions peak to peak.

$$\frac{V_{out}}{V_{out\text{-}max}} = \frac{1}{\sqrt{2}} = 0.707 \approx \frac{5}{7} \textit{ divisions}$$

The frequency is then simply read with a frequency counter or the oscilloscope. **Not by reading the dial of the signal generator**.

Risetime

The risetime indicates how quickly a circuit responds. The risetime is the time it takes a waveform to go from 10% of the voltage range to 90% of the voltage range. This is in response to a square wave and the output voltage must settle to a steady-state voltage (0% and 100%). Most oscilloscopes have dotted lines on the graticule marking the 10% and 90% points to aid in this measurement. Usually these dotted lines assume that 0% is the lowest line of the graticule and 100% is the highest line. The measurement, as shown in figure 15, also includes the risetime of the oscilloscope and the squarewave source.

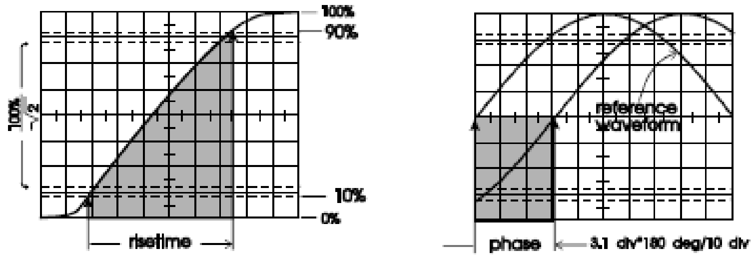


Figure.15 A Risetime and Phase Measurement

Phase

Phase is most accurately measured when the waveform is as large as possible and the difference is measured **at the zero crossings**. Typically the timebase is uncalibrated so that a 180 degree section of the waveform is expanded to the full 10 divisions of the graticule. Then the sign of the phase can be determined by observing more than one period of both waveforms. Both waveforms must be symmetrical about the centre line of the graticule. The angle is determined by: phase = # of divisions * 180 degrees / 10 divisions.

Module 2: Sensors and signal processing
Lecture 1
Sensors and transducers

Measurement is an important subsystem of a mechatronics system. Its main function is to collect the information on system status and to feed it to the micro-processor(s) for controlling the whole system.

Measurement system comprises of sensors, transducers and signal processing devices. Today a wide variety of these elements and devices are available in the market. For a mechatronics system designer it is quite difficult to choose suitable sensors/transducers for the desired application(s). It is therefore essential to learn the principle of working of commonly used sensors/transducers. A detailed consideration of the full range of measurement technologies is, however, out of the scope of this course. Readers are advised to refer "Sensors for mechatronics" by Paul P.L. Regtien, Elsevier, 2012 [2] for more information.

Sensors in manufacturing are basically employed to automatically carry out the production operations as well as process monitoring activities. Sensor technology has the following important advantages in transforming a conventional manufacturing unit into a modern one.

1. Sensors alarm the system operators about the failure of any of the sub units of manufacturing system. It helps operators to reduce the downtime of complete manufacturing system by carrying out the preventative measures.
2. Reduces requirement of skilled and experienced labors.
3. Ultra-precision in product quality can be achieved.

Sensor

It is defined as an element which produces signal relating to the quantity being measured [1]. According to the Instrument Society of America, sensor can be defined as "*A device which provides a usable output in response to a specified measurand.*" Here, the output is usually an 'electrical quantity' and measurand is a 'physical quantity, property or condition which is to be measured'. Thus in the case of, say, a variable inductance displacement element, the quantity being measured is displacement and the sensor transforms an input of displacement into a change in inductance.

The Oscilloscope: Operation and Applications

1. The Oscilloscope

Oscilloscope Operation (X vs Y mode)
An oscilloscope can be used to measure voltage. It does this by measuring the voltage drop across a resistor and in the process draws a small current. The voltage drop is amplified and used to deflect an electron beam in either the X (horizontal) or Y (vertical) axis using an electric field. The electron beam creates a bright dot on the face of the Cathode Ray Tube (CRT) where it hits the phosphorous. The deflection, due to an applied voltage, can be measured with the aid of the calibrated lines on the graticule.
First we will consider the circuitry that amplifies and conditions the voltage to be measured (the "Amp" block in figure 1).

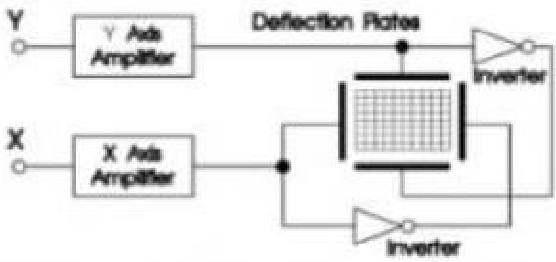


Figure 1. X vs. Y Deflection Block Diagram of the CRT

The deflection of the oscilloscope beam is proportional to the input voltage (after ac or dc coupling). The amount of deflection (Volts/Division) depends upon the setting of the AMPL/DIV control for that channel (see figure 2).

The input signal can be ac or dc coupled. Ac coupling involves adding a series capacitor. This has the effect of blocking (removing) the dc bias and low frequency components of a signal. Dc coupling does not have this problem and therefore allows you to measure voltages right down to 0 Hz. Ac coupling is useful when you are trying to measure a small ac voltage that is "on-top" of a large dc voltage. A typical example is trying to measure the noise of a dc power supply.

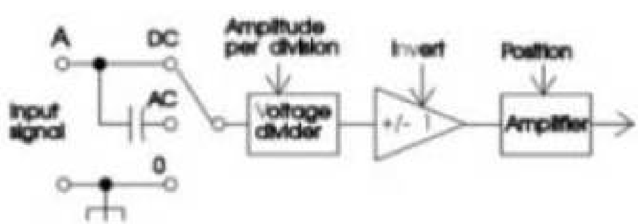


Figure 2. Amplifier Block Diagram

Amplifier Features
AMPL/DIV - This abbreviated name varies but it is generally some short form of amplitude per division. The control is a simple voltage divider (attenuator) which is used to change the sensitivity of the oscilloscope. At a 1 volt/DIV setting, a deflection of one major division on the graticule represents a one volt change at the oscilloscope input.

Calibrated voltage measurements
The small knob within the AMPL/DIV control must be rotated clockwise into its detente

as "*A device which provides a usable output in response to a specified measurand.*" Here, the output is usually an 'electrical quantity' and measurand is a 'physical quantity, property or condition which is to be measured'. Thus in the case of, say, a variable inductance displacement element, the quantity being measured is displacement and the sensor transforms an input of displacement into a change in inductance.

Transducer

It is defined as an element when subjected to some physical change experiences a related change [1] or an element which converts a specified measurand into a usable output by using a transduction principle.

It can also be defined as a device that converts a signal from one form of energy to another form.

A wire of Constantan alloy (copper-nickel 55-45% alloy) can be called as a sensor because variation in mechanical displacement (tension or compression) can be sensed as change in electric resistance. This wire becomes a transducer with appropriate electrodes and input-output mechanism attached to it. Thus we can say that 'sensors are transducers'.

Sensor/transducers specifications

Transducers or measurement systems are not perfect systems. Mechatronics design engineer must know the capability and shortcoming of a transducer or measurement system to properly assess its performance. There are a number of performance related parameters of a transducer or measurement system. These parameters are called as sensor specifications.

Sensor specifications inform the user to the about deviations from the ideal behavior of the sensors. Following are the various specifications of a sensor/transducer system.

1. Range

The range of a sensor indicates the limits between which the input can vary. For example, a thermocouple for the measurement of temperature might have a range of 25-225 °C.

2. Span

The span is difference between the maximum and minimum values of the input. Thus, the above-mentioned thermocouple will have a span of 200 °C.

3. Error

Error is the difference between the result of the measurement and the true value of the quantity being measured. A sensor might give a displacement reading of 29.8 mm, when the actual displacement had been 30 mm, then the error is –0.2 mm.

3. Oscilloscope Applications

Voltage and Time Measurements

Note: The **oscilloscope measures divisions of deflection not voltage or time**. From the divisions of deflection you can calculate the time or voltage.

Differential Measurements

An important application of the oscilloscope is differential measurements. Such measurements are necessary because both vertical channels have one terminal connected to the chassis common (ie single ended). To measure a floating (off ground) voltage you have to use the "invert and add" feature of the oscilloscope. For example, in figure 14, to measure V1:

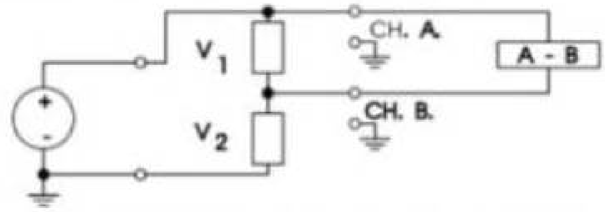


Figure 14. Differential Measurement Example

Channel A measures (V1 + V2) relative to ground while channel B measures V2 relative to ground. By pushing the invert button you negate the voltage displayed on channel B. Then you can add the channels together with the "ADD" display mode. The waveform now displayed is (V1 + V2) + (- V2) = V1.

Bandwidth (-3 dB) Measurement

This measurement is easily done by first finding the maximum gain (max. VOUT / VIN at frequency ω_c) and **adjusting the oscilloscope** so that the sinewave fills seven divisions peak to peak. A -3dB point can be found by increasing and/or decreasing the frequency until the gain is reduced to $\sqrt{2}$. **If the input voltage has remained constant** this will occur when the output voltage is five divisions peak to peak.

$$\frac{V_{out}}{V_{out\text{-}max}} = \frac{1}{\sqrt{2}} = 0.707 = \frac{5}{7} \textit{ divisions}$$

The frequency is then simply read with a frequency counter or the oscilloscope. **Not by reading the dial of the signal generator**.

Risetime

The risetime indicates how quickly a circuit responds. The risetime is the time it takes a waveform to go from 10% of the voltage range to 90% of the voltage range. This is in response to a square wave and the output voltage must settle to a steady-state voltage (0% and 100%). Most oscilloscopes have dotted lines on the graticule marking the 10% and 90% points to aid in this measurement. Usually these dotted lines assume that 0% is the lowest line of the graticule and 100% is the highest line. The measurement, as shown in figure 15, also includes the risetime of the oscilloscope and the squarewave source.

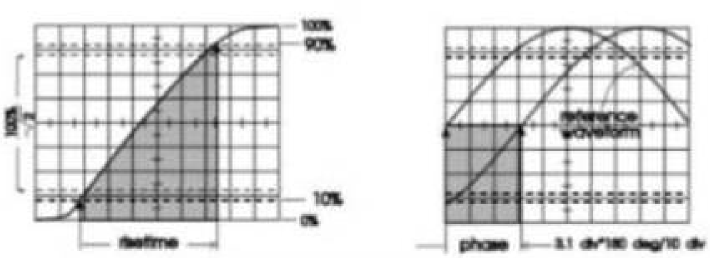


Figure.15 A Risetime and Phase Measurement

Phase

Phase is most accurately measured when the waveform is as large as possible and the difference is measured **at the zero crossings**. Typically the timebase is uncalibrated so that a 180 degree section of the waveform is expanded to the full 10 divisions of the graticule. Then the sign of the phase can be determined by observing more than one period of both waveforms. Both waveforms must be symmetrical about the centre line of the graticule. The angle is determined by: phase = # of divisions * 180 degrees / 10 divisions.

It is wise to consult the user manual for a particular instrument's accuracy specifications.

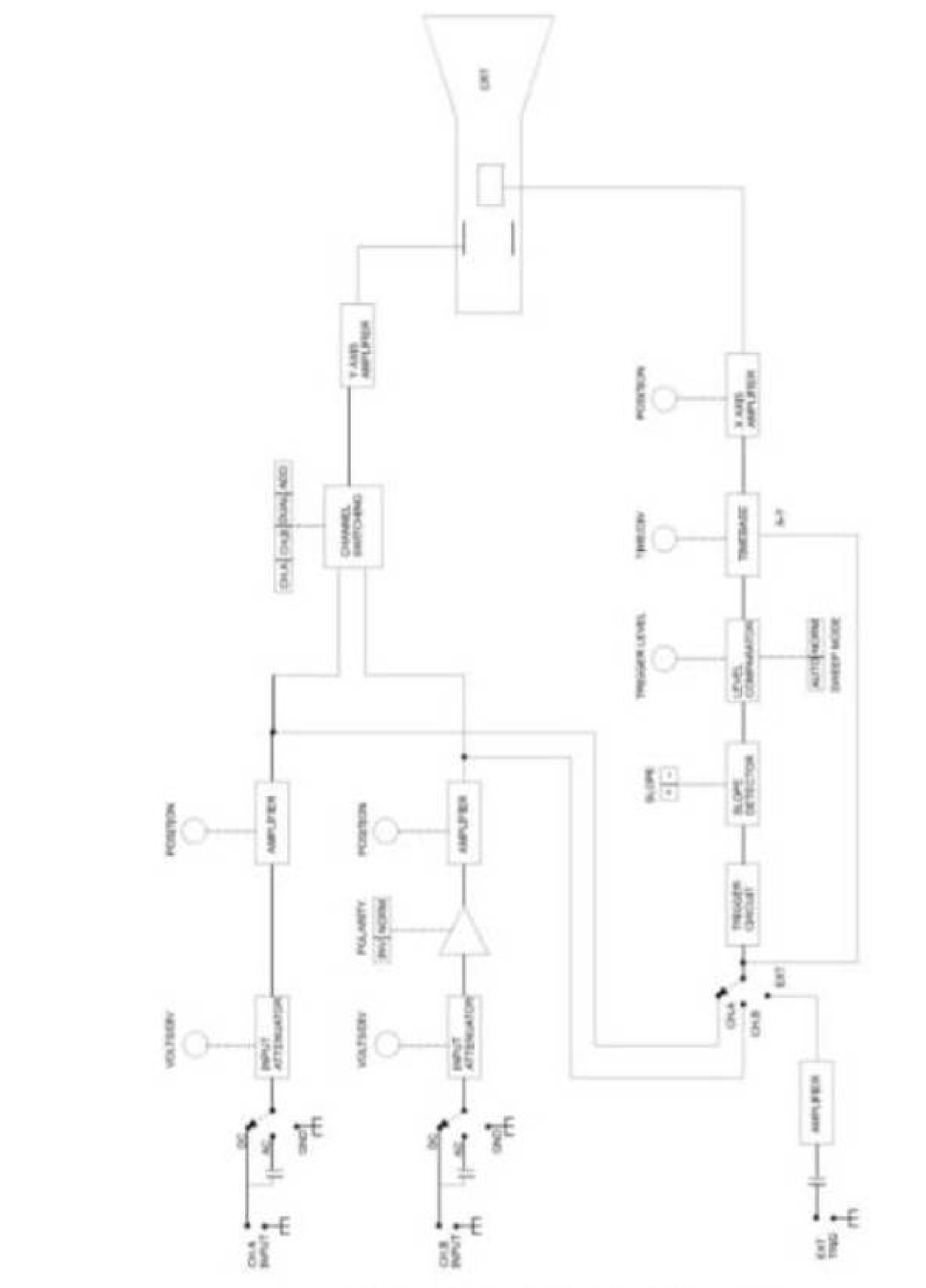


Figure 12. Oscilloscope Block Diagram

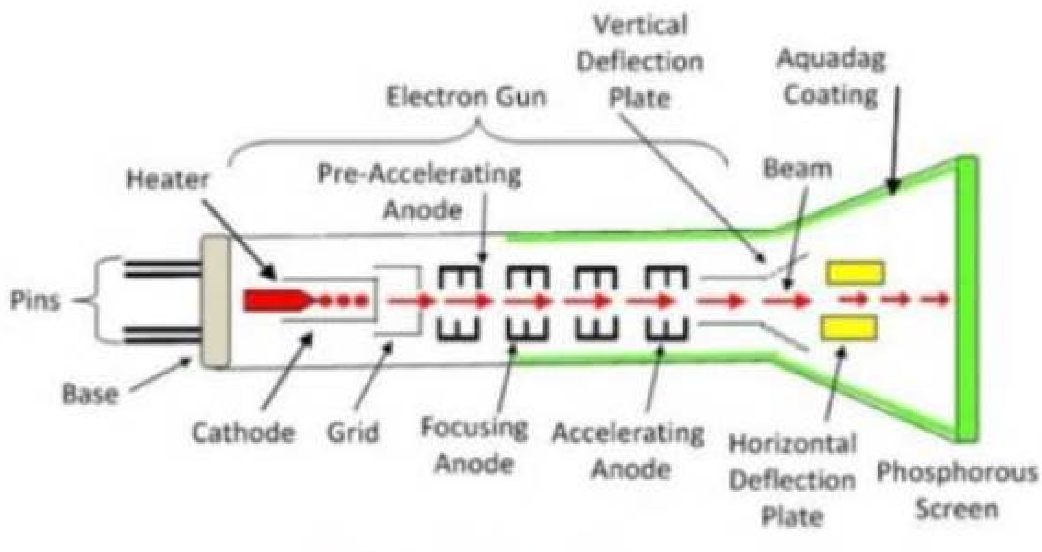
2. Iwatsu Model SS-5702 Oscilloscope

Accuracy

The Iwatsu SS-5702 oscilloscope is used in this laboratory. All measurements on the graticule should be made over as many divisions as possible. For simplicity, assume the Iwatsu SS-5702 oscilloscope's error is ±5% for a measurement on either the vertical or horizontal scales over eight divisions.

Working of CRT

The working of CRT depends on the movement of electrons beams. The electron guns generate sharply focused electrons which are accelerated at high voltage. This high-velocity electron beam when strikes on the fluorescent screen creates luminous spot



Cathode Ray Tube

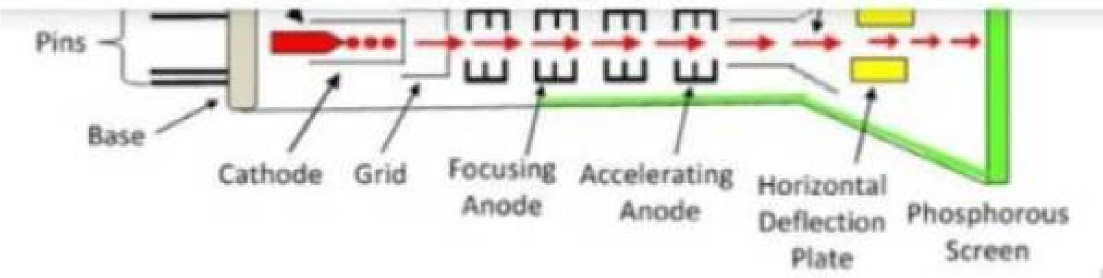
The working parts of a CRT are enclosed in a vacuum glass envelope so that the emitted electron can easily move freely from one end of the tube to the other.

Construction of CRT

The Electrons Gun Assembly, Deflection Plate Assembly, Fluorescent Screen, Glass Envelope, Base are the important parts of the CRT. The electron gun emits the electron beam, and through deflecting plates, it is strikes on the phosphorous screen. The detail explanation of their parts is explained below.

Electrons Gun Assembly

The electron gun is the source of the electron beams. The electron gun has a heater, cathode, grid, pre-accelerating anode, focusing anode and accelerating anode. The electrons are emitted from



Cathode Ray Tube

After exiting from the electron gun, the beam passes through the pairs of electrostatic deflection plate. These plates deflected the beams when the voltage applied across it. The one pair of plate moves the beam upward and the second pair of plate moves the beam from one side to another. The horizontal and vertical movement of the electron are independent of each other, and hence the electron beam positioned anywhere on the screen.

The working parts of a CRT are enclosed in a vacuum glass envelope so that the emitted electron can easily move freely from one end of the tube to the other.

The deflection plate produces the uniform electrostatic field only in the one direction. The electron beam entering into the deflection plates will accelerate only in the one direction, and hence electrons will not move in the other directions.

Screen For CRT

The front of the CRT is called the face plate. The face plate of the CRT is made up of entirely fibre optics which has special characteristics. The internal surface of the faceplate is coated with the phosphor. The phosphorous converts the electrical energy into light energy. The energy level of the phosphorous crystal raises when the electron beams strike on it. This phenomenon is called cathodoluminescence.

the electron beam, and through deflecting plates, it is strikes on the phosphorous screen. The detail explanation of their parts is explained below.

Electrons Gun Assembly

The electron gun is the source of the electron beams. The electron gun has a heater, cathode, grid, pre-accelerating anode, focusing anode and accelerating anode. The electrons are emitted from the highly emitted cathode. The cathode is cylindrical in shape, and at the end of it, the layer of strontium and barium oxide is deposited which emit the high emission of electrons at the end of the tube.

GANDHI ACADEMY OF TECHNOLOGY AND ENGINEERING

GOLANTHARA, BERHAMPUR

LECTURE NOTES

ON

GENERATION,TRANSMISSION AND DISTRIBUTION

For 4th Semester

(As per Syllabus prescribed by SCTE&VT, Odisha)

Prepared By

Mr. Susanta Kumar Sahu

(Lecturer in Electrical Engineering)

Th4. GENERATION TRANSMISSION & DISTRIBUTION

Name of the Course: Diploma in Electrical Engineering			
Course code:		Semester	4^{th}
Total Period:	60	Examination	3 hrs
Theory periods:	4P / week	Internal Assessment :	20
Maximum marks:	100	End Semester Examination:	80

A. RATIONALE :

Power system comprises generation, transmission and distribution. In this subject generation, transmission and distribution, types of generation schemes, transmissionwith transmission loss and efficiencies, different type of sub-stations, different type of distribution schemes, EHV AC and HV DC overhead transmission, underground cable transmission and economic aspects involved are dealt with. Further, types of tariff are briefly included to give brief and overall idea to the students.

B. OBJECTIVES :

After completion of this subject the student will be able to:

1. Different schemes of power generation with their block diagram.
2. Mechanical and electrical design of transmission lines and numerical problems.
3. Types of cables and their methods of laying and testing.
4. Different schemes of distribution with problem solving
5. Different types of sub-stations.
6. Economic aspects of power supply system with problem and type of tariff of electricity.

C. TOPIC WISE DISTRIBUTION OF PERIODS

Sl. No.	Topics	Periods
1.	Generation of electricity	07
2.	Transmission of electric power	05
3.	Over head line	07
4.	Performance of short & medium lines	07
5.	EHV transmission	07
6.	Distribution System	07
7.	Underground cable	06
8.	Economic Aspects	06
9.	Types of tariff	03
10.	Substation	05
	TOTAL	60

D. COURSE CONTENTS IN TERMS OF SPECIFIC OBJECTIVES.

1. GENERATION OF ELECTRICITY

Elementary idea on generation of electricity from Thermal, Hydel, Nuclear,Power station.

Introduction to Solar Power Plant (Photovoltaic cells).

Layout diagram of generating stations.

2. TRANSMISSION OF ELECTRIC POWER

Layout of transmission and distribution scheme.

Voltage Regulation & efficiency of transmission.

State and explain Kelvin's law for economical size of conductor.

Corona and corona loss on transmission lines.

3. OVER HEAD LINES

Types of supports, size and spacing of conductor.

Types of conductor materials.

State types of insulator and cross arms.

Sag in overhead line with support at same level and different level.(approximate formula effect of wind, ice and temperature on sag)

Simple problem on sag.

4. PERFORMANCE OF SHORT & MEDIUM LINES

4.1. Calculation of regulation and efficiency.

5. EHV TRANSMISSION

EHV AC transmission.

5.1..1. Reasons for adoption of EHV AC transmission.5.1..2. Problems involved in EHV transmission.

HV DC transmission.

5.2..1. Advantages and Limitations of HVDC transmission system.

6. DISTRIBUTION SYSTEMS

Introduction to Distribution System.

Connection Schemes of Distribution System: (Radial, Ring Main and Inter connected system)

DC distributions.

Distributor fed at one End.

Distributor fed at both the ends.

Ring distributors.

AC distribution system.

Method of solving AC distribution problem.

Three phase four wire star connected system arrangement.

7. UNDERGROUND CABLES

Cable insulation and classification of cables.

Types of L. T. & H.T. cables with constructional features.

Methods of cable lying.

Localization of cable faults: Murray and Varley loop test for short circuit fault /Earth fault.

8. **ECONOMIC ASPECTS**

Causes of low power factor and methods of improvement of power factor in power system.

Factors affecting the economics of generation: (Define and explain)

Load curves.

Demand factor.

Maximum demand.

Load factor.

Diversity factor.

Plant capacity factor.

Peak load and Base load on power station.

9. **TYPES OF TARIFF**

Desirable characteristic of a tariff.

Explain flat rate, block rate, two part and maximum demand tariff. (SolveProblems)

10. **SUBSTATION**

Layout of LT, HT and EHT substation.

Earthing of Substation, transmission and distribution lines.

Syllabus coverage up to Internal assessment

Chapters: 1, 2, 3, 4 and 5.

Learning Resources:			
Sl.No	Title of the Book	Name of Author	Publisher
1.	Principles of Power System	V. K. Mehta	S Chand
2	A text book of Power System Engineering	A Chakrabarti, M L Soni, P V Gupta, U S Bhatnagar	Dhanpat Rai & Co
3.	A course of electrical power system	S. L. Uppal	Khanna publisher
4.	Power System Engineering	D. P. Kothari, IJ Nagrath	TMH

CHAPTER-1

GENERATION OF ELECTRICITY

Give Elementary idea on generation of electricity from Thermal / Hydel /Nuclear Power station.

Our Universe is consisting by matters and energies. Energy exists in different forms (solar, wind, thermal, geothermal, water energy etc.) in nature but the most important form is the electrical energy which is practically does not exist in nature.

IMPORTANTS OF ELECTRICAL ENERGY:-

Electrical energy is superior to all other forms of energy due to the following reasons:

(i) Convenient form- Electrical energy is a very convenient form of energy. It can be easily convertedinto other forms of energy.

(ii) Easy control-The electrically operated machines have simple and convenient starting, control and operation.

(iii) Greater flexibility- One important reason for preferring electrical energy is the flexibility that it offers. It can be easily transported from one place to another with the help of conductors.

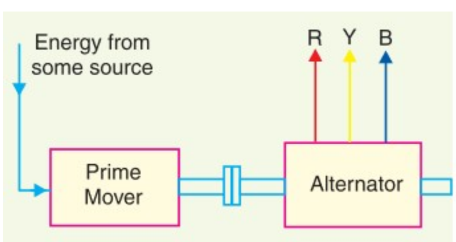
(iv) Cheapness- Electrical energy is much cheaper than other forms of energy.

(v) Cleanliness- Electrical energy is not associated with smoke, fumes or poisonous gases. Therefore, its use ensures cleanliness and healthy conditions.

(vi) High transmission efficiency- The electrical energy can be transmitted conveniently and efficiently from the centers of generation to the consumers with the help of overhead conductors known as transmission lines.

The conversion of energy available in different forms in nature into electrical energy is known as generation of electrical energy.

Energy is available in various forms from different natural sources such as pressure head of water, chemical energy of fuels, nuclear energy of radioactive substances etc. All these forms of energy can be converted into electrical energy by the use of suitable arrangements.



In this arrangement generally an alternator coupled with a prime mover. The prime mover is driven by the energy obtained from various sources existing in nature (such as burning of fuel, pressureof water, force of wind etc.)

SOURCES OF ENERGY:-

Since electrical energy is produced from energy available in various forms in nature, it is desirable to look into the various sources of energy. These sources of energy are : (i) The Sun (ii) The Wind (iii) Water (iv) Fuels (v) Nuclear energy.

(i) The Sun:- The Sun is the primary source of energy. The heat energy radiated by the Sun can be focused over a small area by means of reflectors. This heat can be used to raise steam and electrical energy can be produced with the help of turbine-alternator combination. But it has some limitations like:-

(a) It requires a large area for the generation of even a small amount of electric power

(b) It cannot be used in cloudy days or at night.

(ii) The Wind:- This method can be used where wind flows for a considerable length of time. The wind energy is used to run the wind mill which drives a small generator.

But the drawbacks of this method are:-

(a) Variable output.

(b) Unreliable because of uncertainty about wind pressure and

(c) Power generated is quite small.

(iii) Water:- When water is stored at a suitable place, it possesses potential energy because of the head created. This potential energy of water can be converted into mechanical energy with the help of water turbines. Finally an output (mechanical energy) of water turbine drives the alternator which converts mechanical energy into electrical energy.

This energy considering both Renewable as well as Non-Renewable energy.

(iv) Fuels:- The main sources of energy are fuels Example: solid fuel as coal, liquid fuel as oil and gas fuel as natural gas. The heat energy of these fuels is converted into mechanical energy by suitable prime movers such as steam engines, steam turbines, internal combustion engines etc. and produce an electricity by the help of alternators.

(v) Nuclear energy:- Towards the end of Second World War, it was discovered that large amount of heat energy is liberated by the fission of uranium and other fissionable materials. It is estimated that heat produced by 1 kg of nuclear fuel is equal to that produced by 4500 tons of high grade coal. The heat produced due to nuclear fission can be utilized to produce an electricity.

But there are some difficulties in the use of nuclear energy. The principal ones are (a) high cost of nuclear plant (b) problem of disposal of radioactive waste and dearth of trained personnel to handle the plant.

Generating station:-

The place at which bulk amount of electric power is produced known as generating stations or power plants.

In a generating station essentially used a prime mover coupled to an alternator for the production of electric power. The prime mover (e.g., steam turbine, water turbine etc.) converts energy from some other form into mechanical energy. The alternator converts mechanical energy of the prime mover into electrical energy. The electrical energy produced by the generating station is transmitted and distributed with the help of conductors to various consumers.

Depending upon the form of energy converted into electrical energy, the generating stations are classified as under:

(i) Steam power stations (ii) Hydroelectric power stations

(iii) Diesel power stations (iv) Nuclear power stations.

STEAM POWER STATION (THERMAL POWER STATION)

A generating station which converts heat energy of coal combustion into electrical energy is known as a steam power station.

Schematic arrangement of steam power station:-

The whole arrangement can be divided into the following stages for the sake of simplicity:

1. Coal and ash handling arrangement
2. Steam generating plant
3. Steam turbine
4. Alternator
5. Feed water
6. Cooling arrangement

1. Coal and ash handling plant:- The coal is transported to the power station by road or rail and is stored in the coal storage plant.

From the coal storage plant, coal is delivered to the coal handling plant where it is pulverized (i.e., crushed into small pieces) in order to increase its surface exposure, thus promoting rapid combustion.

2. Steam generating plant: - The steam generating plant consists of a boiler for the production of steam and other auxiliary equipment for the utilization of flue gases.

(i) Boiler- The heat of combustion of coal in the boiler is utilized to convert water into steam at high temperature and pressure. The flue gases from the boiler make their journey through super heater, economizer, air pre-heater and are finally exhausted to atmosphere through the chimney.

(ii) Super heater- The steam produced in the boiler is wet and is passed through a super heater where it is dried and superheated by the flue gases on their way to chimney. Finally the superheated steam from the super heater is fed to steam turbine through the main valve.

(iii) Economizer- An economizer is essentially a feed water heater and derives heat from the flue gases. The economizer extracts a part of heat of flue gases to increase the feed water temperature and fed to the boiler.

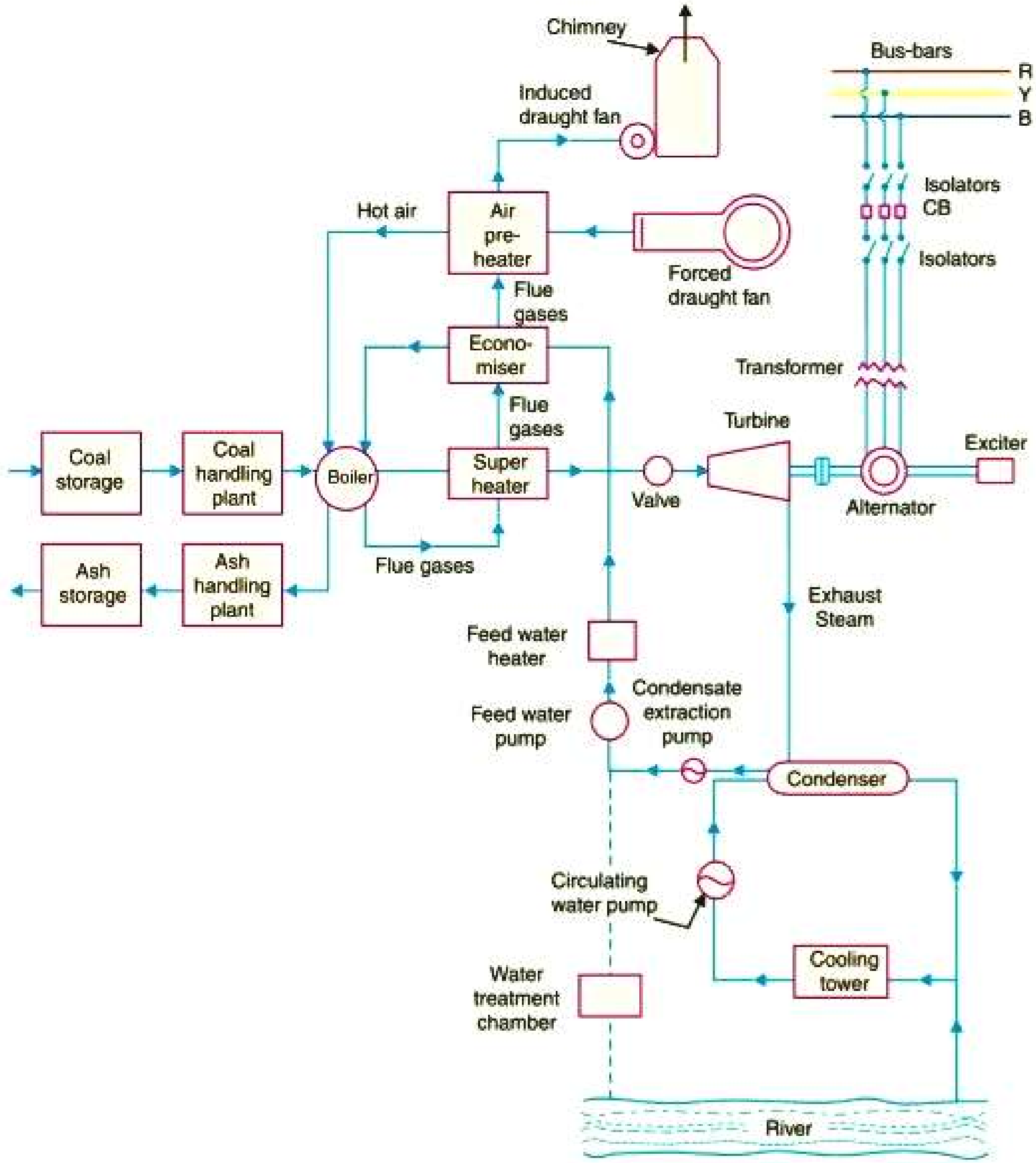
(iv) Air preheater- An air preheater increases the temperature of the air supplied for coal burningby deriving heat from flue gases. Here air is drawn from the atmosphere by a forced draught fan and is passed through air preheater.

3. Steam turbine:- The dry and superheated steam from the super heater is fed to the steam turbine through main valve. The heat energy of steam when passing over the blades of turbine is converted into mechanical energy. After giving heat energy to the turbine, the steam is exhausted to the condenser which condenses the exhausted steam by means of cold water circulation.

4. Alternator: - The steam turbine is coupled to an alternator. The alternator converts mechanical energyof turbine into electrical energy. The electrical output from the alternator is delivered to the bus bars through transformer, circuit breakers and isolators.

5. Feed water: - The condensate from the condenser is used as feed water to the boiler. The feed water on its way to the boiler is heated by water heaters and economizer.

6. Cooling arrangement: - In order to improve the efficiency of the plant, the steam exhausted from the turbine is condensed by means of a condenser. In case the availability of water from the source of supply is not assured throughout the year, cooling towers are used.



Schematic arrangement of Steam Power Station

Advantages of steam power station

(i) The fuel (i.e., coal) used is quite cheap.

(ii) Less initial cost as compared to other generating stations.

(iii) It can be installed at any place irrespective of the existence of coal. The coal can be transported to the site of the plant by rail or road.

(iv) It requires less space as compared to the hydroelectric power station.

(v) The cost of generation is lesser than that of the diesel power station.

Disadvantages of steam power station

(i) It pollutes the atmosphere due to the production of large amount of smoke and fumes.

(ii) It is costlier in running cost as compared to hydroelectric plant.

CHOICE OF SITE FOR STEAM POWER STATION:-

In order to achieve overall economy, the following points should be considered while selecting a site for a steam power station:

(i) Supply of fuel-The steam power station should be located near the coal mines so that transportation cost of fuel is minimum.

(ii) Availability of water- As huge amount of water is required for the condenser; therefore, such a plant should be located at the bank of a river.

(iii) Transportation facilities- A modern steam power station requires the transportation of material and machinery.

(iv) Cost and type of land- The steam power station should be located at a place where land cost is cheap and further extension.

(v) Nearness to load centers- In order to reduce the transmission cost, the plant should be located near the center of the load.

(vi) Distance from populated area- As huge amount of coal is burnt in a steam power station, therefore, smoke and fumes pollute the surrounding area. For which required distance from populated area.
Thermal efficiency- The ratio of heat equivalent of mechanical energy transmitted to the turbine shaftto the heat of combustion of coal is known as thermal efficiency of steam power station.

Thermal efficiency of a modern steam power station is about 30%.
Overall efficiency- The ratio of heat equivalent of electrical output to the heat of combustion of coal isknown as overall efficiency of steam power station.

The overall efficiency of a steam power station is about 29%.

It may be seen that overall efficiency is less than the thermal efficiency. This is expected since some losses (about 1%) occur in the alternator. The following relation exists among the various efficiencies.

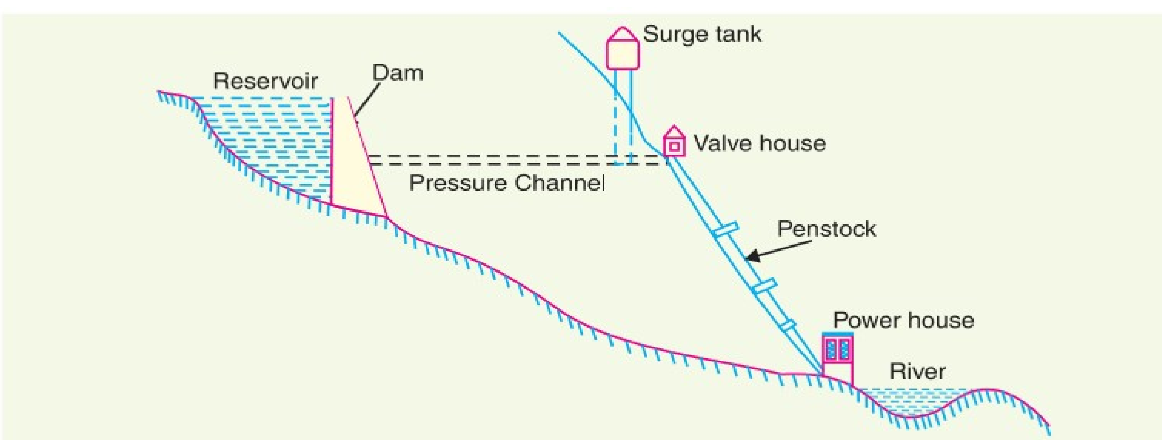
$$\text{Overall efficiency} = \text{Thermal efficiency} \times \text{Electrical efficiency}$$

HYDROELECTRIC POWER STATION

A generating station which utilizes the potential energy of water at a high level for the generation of electrical energy is known as a hydro-electric power station.

Generally hydro-electric power station can be classifying according to head of the dam Like:

(i) Low head hydro-electric power station.(<30mtr)
(ii) Medium head hydro-electric power station.(from 30mtr – 300mtr)
(iii) High head hydro-electric power station.(>300mtr)



Schematic arrangement of a Hydro-electric plant

constituents of a hydro-electric plant:-

The constituents of a hydro-electric plant are (1) hydraulic structures (2) water turbines and (3 electrical equipment.

1. Hydraulic structures- Hydraulic structures in a hydro-electric power station include dam spillways, head works, surge tank, penstock and accessory works.

(i) Dam- A dam is a barrier which stores water and creates water head. Dams are built of concrete or stone masonry, earth or rock fill.

(ii) Spillways- When the river flow exceeds the storage capacity of the reservoir. During heavy rainfall in the catchment area. In order to discharge the surplus water from the storage reservoir into the river on the down-stream side of the dam, spillways are used. Spillways are constructed of concrete piers on the top of the dam.

(iii) Headworks- The headworks consist of the diversion structures at the head of an intake.

(iv) Surge tank- A surge tank is a small reservoir or tank (open at the top) in which water level rises or falls to reduce the pressure swings in the conduit as required. A surge tank is located near the beginning of the conduit. Hence a surge tank overcomes the abnormal pressure in the conduit.

(v) Penstocks- Penstocks are open or closed conduits which carry water from penstock to the power house (turbines). They are generally made of reinforced concrete (<30mtr low head) or steel (for medium and high head).

WORKING:-

Hydro-electric power stations are generally located in hilly areas, where dams can be built conveniently and large water reservoirs can be obtained. In a hydro-electric power station, water head is created by constructing a dam across a river or lake. From the dam, water is led to a water turbine by the help of nozzle through penstock. The water turbine captures the potential energy in the falling water and changes the hydraulic energy into mechanical energy at the turbine shaft. The turbine drives the alternator which converts mechanical energy into electrical energy.

Advantages:-

(i) It requires no fuel as water is used for the generation of electrical energy.

(ii) It is quite neat and clean as no smoke or ash is produced.

(iii) It requires very small running cost because water is the source of energy which is available free of cost.

(iv) It is comparatively simple in construction and requires less maintenance.

(v) It is robust and has a longer life.

(vi) Such plants serve many purposes. In addition to the generation of electrical energy, they also help in irrigation and controlling floods.

Disadvantages:-

(i) It involves high capital cost due to construction of dam.

(ii) There is uncertainty about the availability of huge amount of water due to dependence on weather conditions.

(iii) Skilled and experienced hands are required to build the plant.

(iv) It requires high cost of transmission lines as the plant is located in hilly areas which are quite away from the consumers.

CHOICE OF SITE FOR HYDRO-ELECTRIC POWER STATION:-

In order to achieve overall economy, the following points should be considered while selecting a site for a hydroelectric power station:

(i) Availability of water- Since the primary requirement of a hydro-electric power station is the availability of huge quantity of water.

(ii) Storage of water- it is necessary to store water by constructing a dam in order to ensure the

generation of power throughout the year.

(iii) Transportation facilities- The site selected for a hydro-electric plant should be accessible by rail and road so that necessary equipment and machinery could be easily transported.

(iv) Cost and type of land- The land for the construction of the plant should be available at a reasonable price.

Note:

Generally Water turbines are used to convert the energy of falling water into mechanical energy. The principal types of water turbines are : (i) Impulse turbines (ii) Reaction turbines

(i) **Impulse turbines**. Such turbines are used for high heads. In an impulse turbine, the entire pressure of water is converted into kinetic energy in a nozzle and the velocity of the jet drives the wheel.

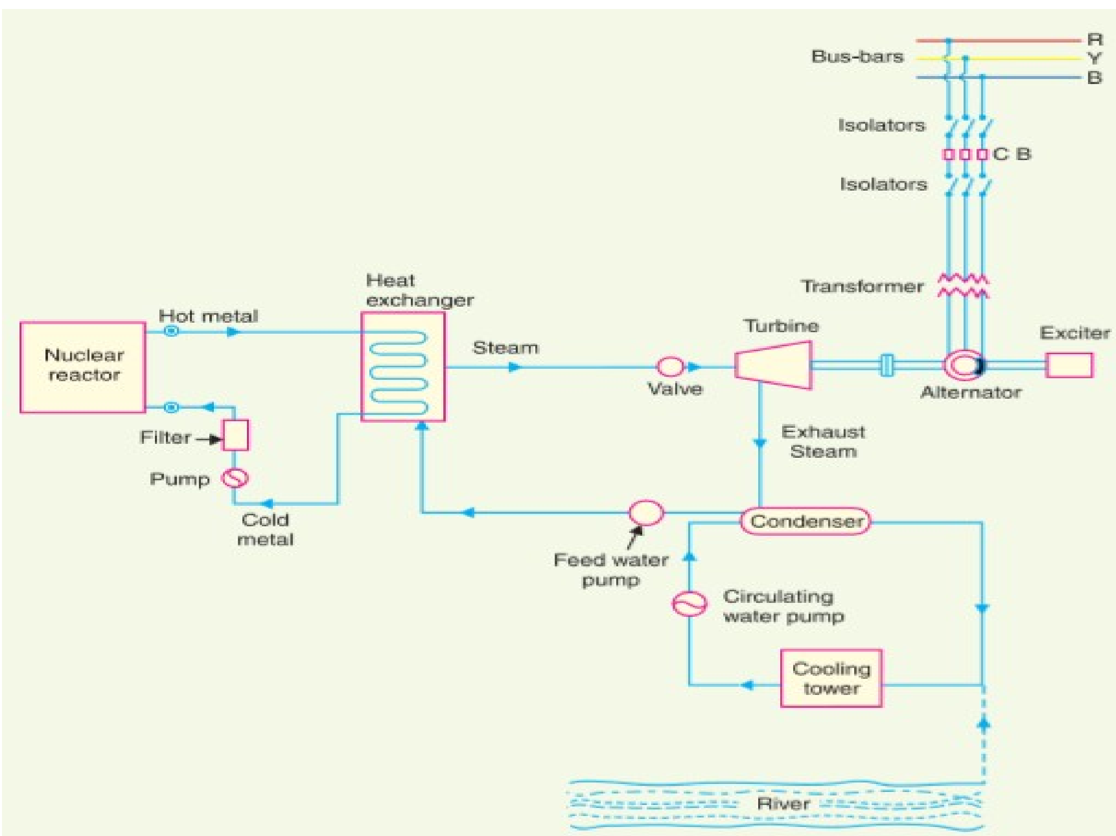
(ii) **Reaction turbines.** It is used for low and medium heads. In a reaction turbine, water enters the runner partly with pressure energy and partly with velocity head. The important types of reaction turbines are : (a) Francis turbines (**used from low to medium head**)
(b) Kaplan turbines (**used only in low head**)

NUCLEAR POWER STATION

A generating station in which nuclear energy is converted into electrical energy is known as a nuclear power station.

In nuclear power station, heavy elements such as Uranium (U^{235}) or Thorium (Th^{232}) are subjected to nuclear fission in a special apparatus known as a reactor. The heat energy thus released is utilized in raising steam at high temperature and pressure. The steam runs the steam turbine which converts steam energy into mechanical energy. The turbine drives the alternator which converts mechanical energy into electrical energy.

It has been found that complete fission of 1 kg of Uranium (U^{235}) can produce as much energy ascan be produced by the burning of 4,500 tons of high grade coal.



Schematic arrangement of Nuclear Power Station

Schematic Arrangement of Nuclear Power Station :-

The schematic arrangement of a nuclear power station is shown in above fig. The whole arrangement can be divided into the following main stages:

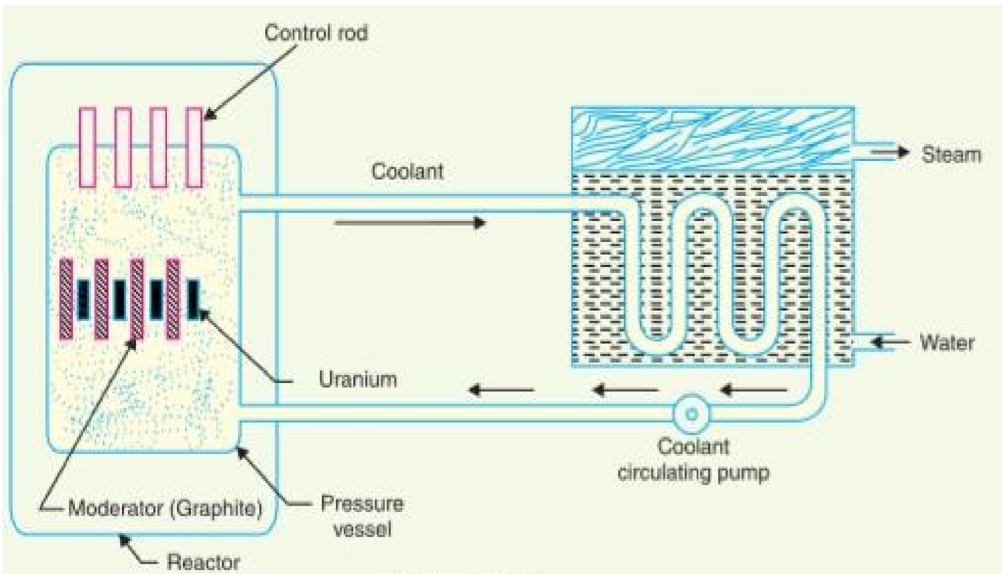
(i) Nuclear reactor (ii) Heat exchanger (iii) Steam turbine (iv) Alternator.

(i) Nuclear reactor-It is an apparatus in which nuclear fuel (U^{235}) is subjected to nuclear fission. It controls the chain reaction that starts once the fission is done. A nuclear reactor is a cylindrical stout pressure vessel and houses fuel rods of Uranium, moderator and control rods as shown in figure.

The moderator consists of graphite rods which enclose the fuel rods. The moderator slows down the neutrons before they bombard the fuel rods.

The control rods are of cadmium and are inserted into the reactor. Cadmium is strong neutron absorber and thus regulates the supply of neutrons for fission.

The heat produced in the reactor is removed by the coolant, generally a sodium metal. The coolant carries the heat to the heat exchanger.



Nuclear Reactor

(ii) Heat exchanger- The coolant gives up heat to the heat exchanger which is utilized in raising the steam. After giving up heat, the coolant is again fed to the reactor.

(iii) Steam turbine- The steam produced in the heat exchanger is led to the steam turbine through a valve. After doing a useful work in the turbine, the steam is exhausted to condenser. The condenser condenses the steam which is fed to the heat exchanger through feed water pump.

(iv) Alternator- The steam turbine drives the alternator which converts mechanical energy into electrical energy. The output from the alternator is delivered to the bus-bars through transformer, circuit breakers and isolators.

Advantages:-

(i) The amount of fuel required is quite small. Therefore, there is a considerable saving in the cost of fuel transportation.

(ii) A nuclear power plant requires less space as compared to any other type of the same size.

(iii) It has low running charges as a small amount of fuel is used for producing bulk electrical energy.

(iv) This type of plant is very economical for producing bulk electric power

Disadvantages:-

(i) **The fuel used is expensive and is difficult to recover.**

(ii) **The capital cost on a nuclear plant is very high as compared to other types of plants.**

(iii) **The erection and commissioning of the plant requires greater technical know-how.**

(iv) **The fission by-products are generally radioactive and may cause a dangerous amount of radioactive pollution.**

(v) **The disposal of the by-products, which are radioactive, is a big problem. They have eitherto be disposed off in a deep trench or in a sea away from sea-shore.**

Selection of Site for Nuclear Power Station:-

(i) **Availability of water. As sufficient water is required for cooling purposes, therefore, the plant site should be located near a river or by sea-side.**

(ii) **Disposal of waste. The waste produced by fission in a nuclear power station is generally radioactive which must be disposed off properly to avoid health hazards.**

(iii) **Distance from populated areas. The site selected for a nuclear power station should be quite away from the populated areas as there is a danger of presence of radioactivity in the atmosphere near the plant.**

(iv)**Transportation facilities. The site selected for a nuclear power station should have adequate facilities in order to transport the heavy equipment during erection and to facilitate the movement of theworkers employed in the plant.**

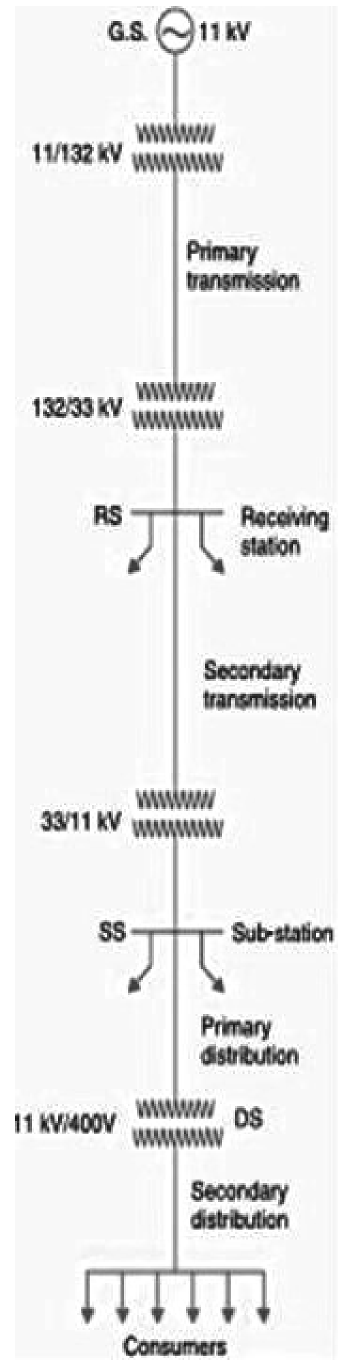
CHAPTER-2

TRANSMISSION OF ELECTRIC POWER

Draw layout of transmission and distribution scheme.

The conveyance of electric power from a power station to consumers' premises is known as electric supply system. An electric supply system consists of three principal components the Generating power station, the transmission lines and the distribution system.

Layout of single line power stage diagram from generating station to consumer end:-



Now-adays, 3-phase, 3-wire a.c. system is universally adopted for generation and transmission of electric power as an economical proposition. However, distribution of electric power is done by 3- phase, 4-wire a.c. system.

(i) Generating Station. G.S. represents the generating station where electric power is produced by 3-phase alternators operating in parallel. The usual generation voltage is 11 kV. For economy in the transmission of electric power, the generation voltage (i.e., 11 kV) is stepped upto 132 kV (or more) at the generating station with the help of 3-phase power transformers and fed to transmission line through tower and conductor.

(ii) Primary transmission. The electric power at 132 kV is transmitted by 3-phase, 3-wire overhead system to the outskirts of the city. This forms the primary transmission. At the receiving station, the voltage is reduced to 33kV by step-down transformers and fed to the substation through secondary transmission.

(iii) Secondary transmission. The secondary transmission line terminates at the sub-station (SS) where voltage is reduced from 33 kV to 11kV, 3-phase, 3-wire. The 11 kV lines run along the important road sides of the city through primary distribution.

(iv) Primary distribution. The electric power from primary distribution line (11 kV) is delivered to distribution sub-stations (DS). These sub-stations are located near the consumers' localities and step down the voltage to 400 V.

(v) Secondary distribution. Finally by 3-phase, 4-wire system it is fed to all consumers by secondary distribution system (consists of feeders, distributors and service mains). The voltage between any two phases is 400 V and between any phase and neutral is 230 V.

Explain voltage Regulation & efficiency of transmission.

Important term :-

(i) Voltage regulation. When a transmission line is carrying current, there is a voltage drop in the line due to resistance and inductance of the line. The result is that receiving end voltage (V_R) of the line is generally less than the sending end voltage (V_S). This voltage drop ($V_S - V_R$) in the line is expressed as a percentage of receiving end voltage V_R and is called voltage regulation.

"The difference in voltage at the receiving end of a transmission line between conditions of no load and full load is called voltage regulation and is expressed as a percentage of the receiving end voltage"

Mathematically,

$$\% \text{ age Voltage regulation } = \frac{V_S - V_R}{V_R} \times 100$$

(ii) Transmission efficiency. The power obtained at the receiving end of a transmission line is generally less than the sending end power due to losses in the line resistance.

"The ratio of receiving end power to the sending end power of a transmission line is known as the transmission efficiency of the line"

$$\% \text{ age Transmission efficiency, } \eta_T = \frac{\text{Receiving end power}}{\text{Sending end power}} \times 100$$

$$= \frac{V_R I_R \cos \varphi_R}{V_S I_S \cos \varphi_S} \times 100$$

State and explain Kelvin's law for economical size of conductor.

Economics of Power Transmission:-

The following two fundamental economic principles which closely influence the electrical design transmission line will be discussed:

(i) Economic choice of conductor size.

(ii) Economic choice of transmission voltage.

Economic choice of conductor size

The cost of conductor material is generally a very considerable part of the total cost of a transmission line. Therefore, the determination of proper size of conductor for the line is more importance. The most economical area of conductor is that for which the total annual cost of transmission line is minimum. This is known as Kelvin's Law after Lord Kelvin who first stated it in 1881. The total annual cost of transmission line can be divided broadly into two parts as: - annual charge on capital outlay and annual cost of energy wasted in the conductor.

Kelvin's law:-

Kelvin's Law can also be stated in another way i.e. "the most economical area of conductor is that for which the variable part of annual charge is equal to the cost of energy wasted per year."

(i) Annual charge on capital outlay. It involves both interest and depreciation on the capital cost of complete installation of transmission line. In case of overhead system, it will be the annual interest and depreciation on the capital cost of conductors, supports and insulators and the cost of theirerection. Now, for an overhead line, insulator cost is constant, the conductor cost is proportional to the area of X-section and the cost of supports and their erection is partly constant and partly proportional to area of X-section of the conductor. Therefore, annual charge on an overhead† transmission line canbe expressed as :

$$\text{Annual charge } = P_1 + P_2 a \quad \text{............ (i)}$$

where P1 and P2 are constants and a is the area of X -section of the conductor.

(ii) Annual cost of energy wasted. This is on account of energy lost mainly in the conductor due to I2R losses. Assuming a constant current in the conductor throughout the year, the energy lost inthe conductor is proportional to resistance. As resistance is inversely proportional to the area of X-section of the conductor, therefore, the energy lost in the conductor is inversely proportional to area of X-section. Thus, the annual cost of energy wasted in an overhead transmission line can be expressed as :

$$\text{Annual cost of energy wasted} = P_3/a \qquad \text{............(ii)}$$

where P_3 is a constant.

$$\text{Total annual cost, } C = \text{exp. } (i) + \text{exp. } (ii)$$
$$= (P_1 + P_2\, a) + P_3/a$$
$$\therefore \qquad C = P_1 + P_2\, a + P_3/a$$

In exp. (*iii*), only area of X-section a is variable. Therefore, the total a
line will be minimum if differentiation of C w.r.t. a is zero *i.e.*

$$\frac{d}{da}(C) = 0$$

$$\text{or} \qquad \frac{d}{da}(P_1 + P_2\, a + P_3/a) = 0$$

$$\text{or} \qquad P_2 - \frac{P_3}{a^2} = 0$$

$$\text{or} \qquad P_2 = P_3/a^2$$

$$\text{or} \qquad P_2\, a = \frac{P_3}{a}$$

i.e. Variable part of annual charge = Annual cost of energy wasted

It is clear that the most economical area of conductor is that for which the variable part of annua charge is equal to the cost of energy losses per year.

Graphical illustration of Kelvin's law. Kelvin's law can also be illustrated graphically by plotting annual cost against X-sectional area 'a' of the conductor as shown in Fig. 7.28. In the diagram, the straight line (1) shows the relation between the annual charge (*i.e.*, $P_1 + P_2 a$) and the area of X-section a of the conductor. Similarly, the rectangular hyperbola (2) gives the relation between annual cost of energy wasted and X-sectional area a. By adding the ordinates of curves (1) and (2), the curve (3) is obtained. This latter curve shows the relation between total annual cost ($P_1 + P_2 a + P_3/a$) of transmission line and area of X-section a. The lowest point on the curve (*i.e.*, point P) represents the most economical area of X-section.

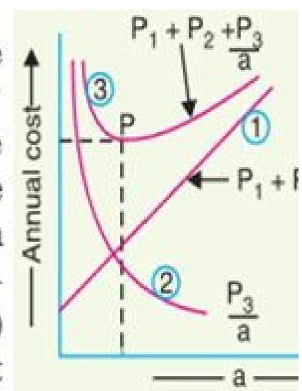


Fig. 7.28

Limitations of Kelvin's law. Although theoretically Kelvin's law holds good, there is often considerable difficulty in applying it to a proposed scheme of power transmission. In practice, the limitations of this law are :

(i) It is not easy to estimate the energy loss in the line without actual load curves, which are not available at the time of estimation.

(ii) The assumption that annual cost on account of interest and depreciation on the capital outlay is in the form P1 + P2a is strictly speaking not true. For instance, in cables neither the cost of cable dielectric and sheath nor the cost of laying vary in this manner.

(iii) This law does not take into account several physical factors like safe current density, mechanical strength, corona loss etc.

(iv) The conductor size determined by this law may not always be practicable one because it may be too small for the safe carrying of necessary current.

(v) Interest and depreciation on the capital outlay cannot be determined accurately.

Explain corona and corona loss on transmission lines.

CORONA:-

The phenomenon of violet glow, hissing noise and production of ozone gas in an overhead transmission line is known as corona.

Theory of corona formation:-

Some ionization is always present in air due to cosmic rays, ultraviolet radiations and radioactivity. Therefore, under normal conditions, the air around the conductors contains some ionized particles (i.e., free electrons and +ve ions) and neutral molecules. When p.d. is applied between the conductors, potential gradient is set up in the air which will have maximum value at the conductor surfaces. Under the influence of potential gradient, the existing free electrons acquire greater velocities.

When the potential gradient at the conductor surface reaches about 30 kV per cm (max. value), the velocity acquired by the free electrons is sufficient to strike a neutral molecule with enough force. This produces another ion and one or more free electrons. Thus, the process of ionisation is cummulative. The result of this ionisation is that either corona is formed.

FACTORS AFFECTING CORONA:-

The following are the factors upon which corona depends :

(i) Atmosphere. As corona is formed due to ionization of air surrounding the conductors, therefore, itis affected by the physical state of atmosphere. In the stormy weather, the number of ions is more than normal and as such corona occurs at much less voltage as compared with fair weather.

(ii) Conductor size. The corona effect depends upon the shape and conditions of the conductors. The rough and irregular surface will give rise to more corona. Thus a stranded conductor has irregular surface and hence gives rise to more corona that a solid conductor.

(iii) Spacing between conductors. If the spacing between the conductors is made very large as compared to their diameters, there may not be any corona effect. It is because larger distance between conductors reduces the electro-static stresses at the conductor surface, thus avoiding corona formation.

(iv) Line voltage. The line voltage greatly affects corona. If it is low, no corona is formed. However, if the line voltage is high, due to produce high electro-static stress corona is formed.

IMPORTANT TERMS USED IN CORONA

The phenomenon of corona plays an important role in the design of an overhead transmission line. Therefore, it is profitable to consider the following terms much used in the analysis of corona

effects:

(i) Critical disruptive voltage. "It is the minimum phase-neutral voltage at which corona occurs". Consider two conductors of radii r cm and spaced d cm apart. If V is the phase-neutral potential, then potential gradient at the conductor surface is given by:

$$g = \frac{V}{r \log_e \frac{d}{r}} \text{ volts / cm}$$

In order that corona is formed, the value of g must be made equal to the breakdown strength of air. The breakdown strength of air at 76 cm pressure and temperature of 25ºC is 30 kV/cm (max) or 21·2 kV/cm (r.m.s.) and is denoted by go. If VC is the phase-neutral potential required under these conditions, then,

$$g_0 = \frac{V_c}{r \log_e \frac{d}{r}}$$

where g_0 = breakdown strength of air at 76 cm of mercury and 25°C

$= 30$ kV/cm (*max*) or $21{\cdot}2$ kV/cm (*r.m.s.*)

$$\therefore \text{ Critical disruptive voltage, } V_c = g_0\, r \log_e \frac{d}{r}$$

The above expression for disruptive voltage is under standard conditions i.e., at 76 cm of Hg and 25ºC. However, if these conditions vary, the air density also changes, thus altering the value of g_0. The value of g_0 is directly proportional to air density. Thus the breakdown strength of air at a barometric pressure of b cm of mercury and temperature of tºC becomes δ g_0 where

$$\delta = \text{air density factor} = \frac{3 \cdot 92 b}{273 + t}$$

Under standard conditions, the value of $\delta = 1$.

$$\therefore \text{ Critical disruptive voltage, } V_c = g_o\, \delta\, r \log_e \frac{d}{r}$$

Correction must also be made for the surface condition of the conductor. This is accounted for by multiplying the above expression by irregularity factor m_0.

$$\therefore \text{ Critical disruptive voltage, } V_c = m_o\, g_o\, \delta\, r \log_e \frac{d}{r} \text{ kV/phase}$$

where m_o = 1 for polished conductors

= 0·98 to 0·92 for dirty conductors

= 0·87 to 0·8 for stranded conductors

(ii) Visual critical voltage. "It is the minimum phase-neutral voltage at which corona glow appears all along the line conductors".

It has been seen that in case of parallel conductors, the corona glow does not begin at the disruptive voltage V_c but at a higher voltage V_v called visual critical voltage. The phase-neutral effective value of visual critical voltage is given by the following empirical formula :

$$V_v = m_v g_o\, \delta\, r \left(1 + \frac{0 \cdot 3}{\sqrt{\delta r}}\right) \log_e \frac{d}{r} \text{ kV/phase}$$

where mv is another irregularity factor having a value of 1·0 for polished conductors and 0·72 to 0·82 for rough conductors.

(iii) Power loss due to corona. Formation of corona is always accompanied by energy loss which is dissipated in the form of light, heat, sound and chemical action. When disruptive voltage is exceeded, the power loss due to corona is given by :

$$P = 242 \cdot 2 \left(\frac{f+25}{\delta}\right)\sqrt{\frac{r}{d}}\,(V - V_c)^2 \times 10^{-5} \text{ kW / km / phase}$$

where
f = supply frequency in Hz
V = phase-neutral voltage ($r.m.s.$)
V_c = disruptive voltage ($r.m.s.$) per phase

Advantages and Disadvantages of Corona:-

Advantages

(i) Due to corona formation, the air surrounding the conductor becomes conducting and hence virtual diameter of the conductor is increased. The increased diameter reduces the electrostatic stresses between the conductors.

(ii) Corona reduces the effects of transients produced by surges.

Disadvantages

(i) Corona is accompanied by a loss of energy. This affects the transmission efficiency of the line.

(ii) Ozone is produced by corona and may cause corrosion of the conductor due to chemical action.

(iii) The current drawn by the line due to corona is non-sinusoidal and hence non-sinusoidal voltage drop occurs in the line. This may cause inductive interference with neighboring communication lines.

Methods of Reducing Corona Effect

In general corona effects are observed at a working voltage of 33 kV or above. The corona effects can be reduced by the following methods :

(i) By increasing conductor size. By increasing conductor size, the voltage at which corona occurs is raised and hence corona effects are considerably reduced. This is one of the reasons that ACSR conductors which have a larger cross-sectional area are used in transmission lines.

(ii) By increasing conductor spacing. By increasing the spacing between conductors, the voltage at which corona occurs is raised and hence corona effects can be eliminated. However, spacing cannot be increased too much otherwise the cost of supporting structure (e.g., bigger cross arms and supports) may increase to a considerable extent.

CHAPTER-3

OVERHEAD LINES

State types of supports, size and spacing of conductor.

Electric power can be transmitted or distributed either by means of underground cables or by overhead lines. An overhead line is subjected to uncertain weather conditions and other external interferences. This calls for the use of proper mechanical factors of safety in order to ensure the continuity of operation in the line. In general, the strength of the line should be such so as to provide against the worst probable weather conditions. In this chapter, we shall focus our attention on the various aspects of mechanical design of overhead lines.

Main Components of Overhead Lines :-

(i) Conductors which carry electric power from the sending end station to the receiving end station.

(ii) Supports which may be poles or towers and keep the conductors at a suitable level above the ground.

(iii) Insulators which are attached to supports and insulate the conductors from the ground.

(iv) Cross arms which provide support to the insulators.

(v) Miscellaneous items such as phase plates, danger plates, lightning arrestors, anti-climbing wires etc.

LINE SUPPORT

The supporting structures for overhead line conductors are various types of poles and towers called line supports. In general, the line supports should have the following properties :

(i) High mechanical strength to withstand the weight of conductors and wind loads etc.
(ii) Light in weight without the loss of mechanical strength.
(iii) Cheap in cost and economical to maintain.
(iv) Longer life.
(v) Easy accessibility of conductors for maintenance.

The line supports used for transmission and distribution of electric power are of various types including wooden poles, steel poles, R.C.C. poles and lattice steel towers. The choice of supporting structure for a particular case depends upon the line span, X-sectional area, line voltage, cost and local conditions.

1. Wooden poles. These are made of seasoned wood (Sal or chirr) and are suitable for lines of moderate X-sectional area and of relatively shorter spans, say up to 50 meters. Such supports are cheap, easily available, provide insulating properties and, therefore, are widely used for distribution purposes in rural areas as an economical proposition. The wooden poles generally tend to rot below the ground level, causing foundation failure. In order to prevent this, the portion of the pole below the ground level is impregnated with preservative compounds like creosote oil. Double pole structures of the 'A' or 'H' type are often used to obtain a higher transverse strength than could be economically provided by means of single poles.

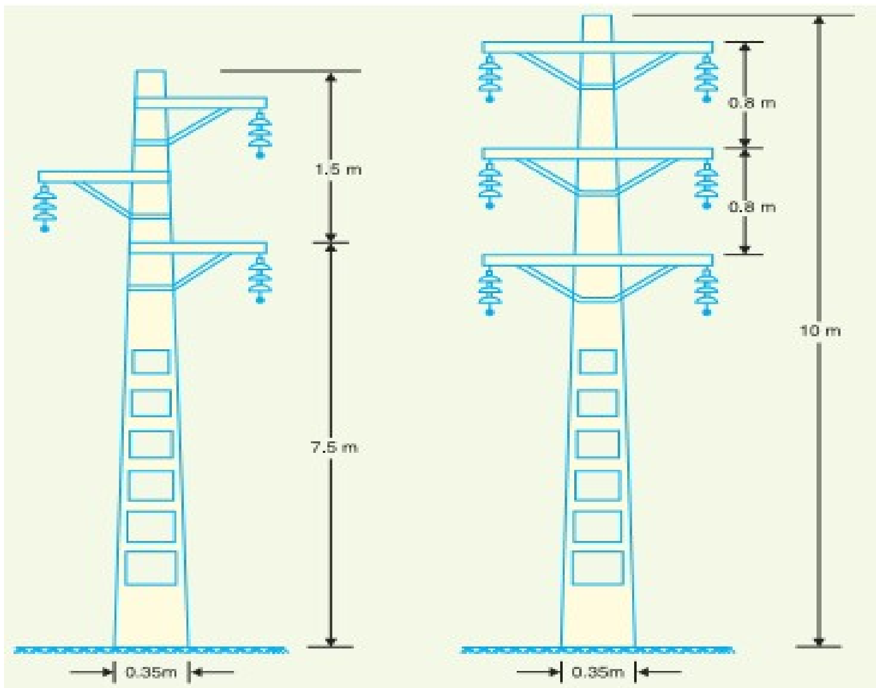
The main objections to wooden supports are : (i) tendency to rot below the ground level (ii) comparatively smaller life (20-25 years) (iii) cannot be used for voltages higher than 20 kV (iv) less mechanical strength and (v) require periodical inspection.

2. Steel poles. The steel poles are often used as a substitute for wooden poles. They possess greater mechanical strength, longer life and permit longer spans to be used. Such poles are generally used for

distribution purposes in the cities. This type of supports need to be galvanized or painted in order to prolong its life. The steel poles are of three types viz., (i) rail poles (ii) tubular poles and (iii) rolled steel joints.

3. RCC poles. The reinforced concrete poles have become very popular as line supports in recent years. They have greater mechanical strength, longer life and permit longer spans than steel poles. Moreover, they give good outlook, require little maintenance and have good insulating properties. The holes in the poles facilitate the climbing of poles and at the same time reduce the weight of line supports. The main difficulty with the use of these poles is the high cost of transport owing to their heavy weight. Therefore, such poles are often manufactured at the site in order to avoid heavy cost of transportation.

4. Steel towers. In practice, wooden, steel and reinforced concrete poles are used for distribution purposes at low voltages, say up to 11 kV. However, for long distance transmission at higher voltage, steel towers are invariably employed. Steel towers have greater mechanical strength, longer life, can withstand most severe climatic conditions and permit the use of longer spans. The risk of interrupted service due to broken or punctured insulation is considerably reduced owing to longer spans. Tower footings are usually grounded by driving rods into the earth. This minimizes the lightning troubles as each tower acts as a lightning conductor.



Types of conductor materials.

Conductor Materials:-

The conductor material used for transmission and distribution of electric power should have the following properties:

(i) High electrical conductivity or low resistivity.

(ii) High tensile strength in order to withstand mechanical stresses.

(iii) low cost so that it can be used for long distances.

(iv) Low specific gravity so that weight per unit volume is small.

Commonly used conductor materials. The most commonly used conductor materials for overhead lines are copper, aluminum, steel-cored aluminum, galvanized steel and cadmiumcopper. The choice of a particular material will depend upon the cost, the required electrical and mechanical properties and the local conditions.

All conductors used for overhead lines are preferably stranded (ACSR) order to increase the flexibility.In stranded conductors, there is generally one central wire and round this, successive layers of wires containing 6, 12, 18, 24 ...... wires. Thus, if there are n layers, the total number of individual wires is 3n (n + 1) + 1.

1.Copper. Copper is an ideal material for overhead lines owing to its high electrical conductivity and greater tensile strength. It is always used in the hard drawn form as stranded conductor. Although hard drawing decreases the electrical conductivity slightly yet it increases the tensile strength considerably. This leads to two advantages. Firstly, smaller X-sectional area of conductor is required and secondly, the area offered by the conductor to wind loads is reduced. However, due to its higher cost and non-availability, it is rarely used for these purposes. Now-a-days the trend is to use aluminum in place of copper.

2. Aluminum. Aluminum is cheap and light as compared to copper but it has much smaller conductivity and tensile strength. The relative comparison of the two materials is briefed below:

(i) The conductivity of aluminum is 60% that of copper. For the same resistance, the diameter of aluminum conductor is about 1·26 times the diameter of copper conductor.

(ii) The specific gravity of aluminum (2·71 gm/cc) is lower than that of copper (8·9 gm/cc). Therefore,an aluminum conductor has almost one-half the weight of equivalent copper conductor.

(iii) Aluminum conductor being light is liable to greater swings and hence larger cross-arms are required.

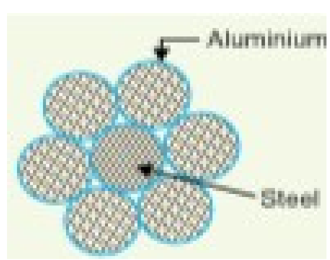
3. Steel cored aluminum. Due to low tensile strength, aluminum conductors produce greater sag. In order to increase the tensile strength, the aluminum conductor is reinforced with a core of galvanized steel wires. The *composite conductor thus obtained is known as steel cored aluminum and is abbreviated as A.C.S.R. (aluminum conductor steel reinforced). The steel cored aluminum conductors have the following advantages :

(i) The reinforcement with steel increases the tensile strength but at the same time keeps the composite conductor light. Therefore, steel cored aluminum conductors will produce smaller sag and hence longerspans can be used.

(ii) Due to smaller sag with steel cored aluminum conductors, towers of smaller heights can be used.

4. Galvanized steel. Steel has very high tensile strength. Therefore, galvanized steel conductors can be used for extremely long spans or for short line sections exposed to abnormally high stresses due to climatic conditions. They have been found very suitable in rural areas where cheapness is the main consideration. Due to poor conductivity and high resistance of steel, such conductors are not suitable for transmitting large power over a long distance.

5. Cadmium copper. The conductor material now being employed in certain cases is copper alloyed with cadmium. An addition of 1% or 2% cadmium to copper increases the tensile strength by about 50% and the conductivity is only reduced by 15% below that of pure copper. Therefore, cadmium copper conductor can be useful for exceptionally long spans.



State types of insulator and cross arms.

The overhead line conductors should be supported on the poles or towers in such a way that currents from conductors do not flow to earth through supports i.e., line conductors must be properly insulated from supports. This is achieved by securing line conductors to supports with the help of insulators. The insulators provide necessary insulation between line conductors and supports and thus prevent any leakage current from conductors to earth. In general, the insulators should have the following desirable properties :

(i) High mechanical strength in order to withstand conductor load, wind load etc.

(ii) High electrical resistance of insulator material in order to avoid leakage currents to earth.

(iii) High relative permittivity of insulator material in order that dielectric strength is high.

(iv) The insulator material should be non-porous; free from impurities and cracks otherwise the permittivity will be lowered.

(iv) High ratio of puncture strength to flashover.

The most commonly used material for insulators of overhead line is porcelain but glass, steatite and special composition materials are also used to a limited extent. Porcelain is produced by firing ata high temperature a mixture of kaolin, feldspar and quartz. It is stronger mechanically than glass, gives less trouble from leakage and is less affected by changes of temperature.

Types of Insulators

There are several types of insulators but the most commonly used are pin type, suspension type, straininsulator and shackle insulator.

Pin type insulators. The part section of a pin type insulator is shown in Fig. 8.5 (i). As the name suggests, the pin type insulator is secured to the cross-arm on the pole. There is a groove on the upperend of the insulator for housing the conductor. The conductor passes through this groove and is bound by the annealed wire of the same material as the conductor Pin type insulators are used for transmissionand distribution of electric power at voltages up to 33 kV. Beyond operating voltage of 33 kV, the pintype insulators become too bulky and hence uneconomical.

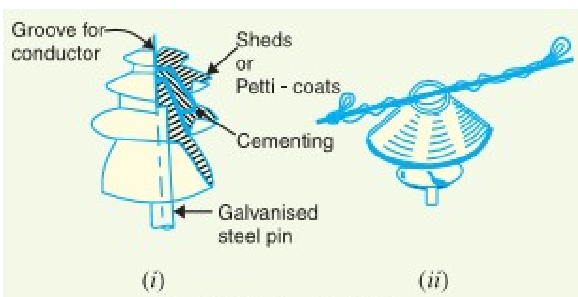


Fig. 8.5. Pin-type insulator

Pin type insulator

Causes of insulator failure.

Insulators are required to withstand both mechanical and electrical stresses. The latter type is primarily due to line voltage and may cause the breakdown of the insulator. The electrical breakdown of the insulator can occur either by flash-over or puncture. In flashover, an arc occurs between the line conductor and insulator pin (i.e., earth) and the discharge jumps across the *air gaps, following shortest distance. Fig. 8.6 shows the arcing distance (i.e. a + b + c) for the insulator.

In case of flash-over, the insulator will continue to act in its proper capacity unless extreme heat produced by the arc destroys the insulator.

In case of puncture, the discharge occurs from conductor to pin through the body of the insulator. When such breakdown is involved, the insulator is permanently destroyed due to excessive heat. In practice, sufficient thickness of porcelain is provided in the insulator to avoid puncture by the line voltage. The ratio of puncture strength to flashover voltage is known as safety factor i.e.,

$$\text{Safety factor of insulator} = \frac{\text{Puncture strength}}{\text{Flash - over voltage}}$$

It is desirable that the value of safety factor is high so that flash-over takes place before the insulator gets punctured. For pin type insulators, the value of safety factor is about 10.

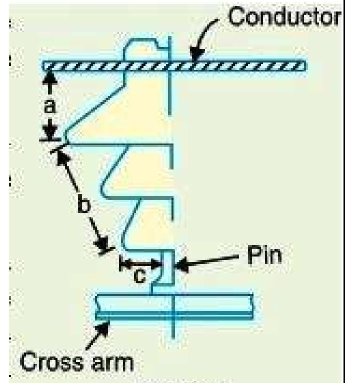


Fig. 8.6

Suspension type insulators. The cost of pin type insulator increases rapidly as the working voltage isincreased. Therefore, this type of insulator is not economical beyond 33 kV. For high voltages (>33 kV), it is a usual practice to use suspension type insulators shown in Fig. 8.7. They consist of a numberof porcelain discs connected in series by metal links in the form of a string. The conductor is suspendedat the bottom end of this string while the other end of the string is secured to the cross-arm of the tower.Each unit or disc is designed for low voltage, say 11 kV. The number of discs in series would obviously depend upon the working voltage. For instance, if the working voltage is 66 kV, then six discs in serieswill be provided on the string.

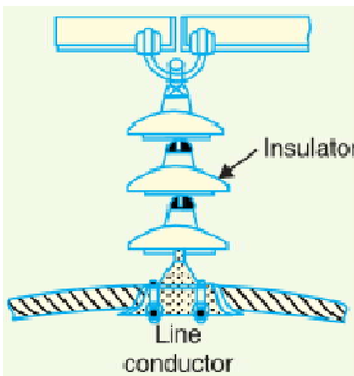


Fig. 8.7


Suspension insulator

Advantages

(i) Suspension type insulators are cheaper than pin type insulators for voltages beyond 33 kV.

(ii) Each unit or disc of suspension type insulator is designed for low voltage, usually 11 kV. Depending upon the working voltage, the desired number of discs can be connected in series.

(iii) If anyone disc is damaged, the whole string does not become useless because the damageddisc can be replaced by the sound one.

(iv) The suspension arrangement provides greater flexibility to the line.

(v) In case of increased demand on the transmission line, it is found more satisfactory to supply the greater demand by raising the line voltage than to provide another set of conductors.

(vi) The suspension type insulators are generally used with steel towers. As the conductors run below the earthed cross-arm of the tower, therefore, this arrangement provides partial protection from lightning.

Strain insulators. When there is a dead end of the line or there is corner or sharp curve, the line is subjected to greater tension. In order to relieve the line of excessive tension, strain insulators are used. For low voltage lines (< 11 kV), shackle insulators are used as strain insulators. However, for high voltage transmission lines, strain insulator consists of an assembly of suspension insulators as shown in Fig. 8.8. The discs of strain insulators are used in the vertical plane. When the tension in lines is exceedingly high, as at long river spans, two or more strings are used in parallel.

Shackle insulators. In early days, the shackle insulators were used as strain insulators. But now a day, they are frequently used for low voltage distribution lines. Such insulators can be used either in a horizontal position or in a vertical position. They can be directly fixed to the pole with a bolt or to the cross arm. Fig. 8.9 shows a shackle insulator fixed to the pole. The conductor in the groove is fixed with a soft binding wire.

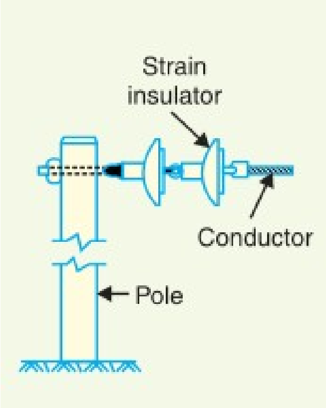


Fig. 8.8. Strain insulator.

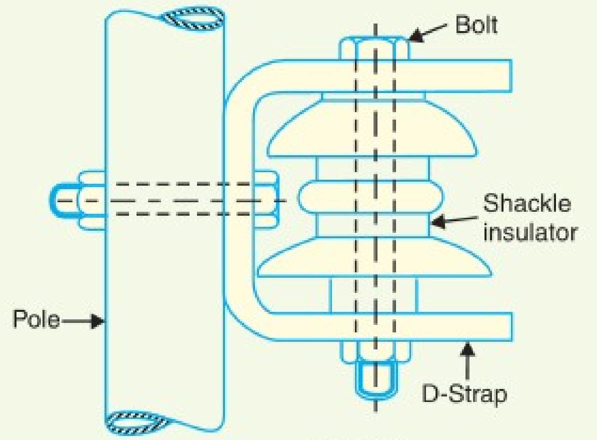


Fig. 8.9

Potential Distribution Over Suspension Insulator String:-

A string of suspension insulators consists of a number of porcelain discs connected in series through metallic links. Fig. 8.10 (i) shows 3-disc string of suspension insulators. The porcelain portion of each disc is in between two metal links. Therefore, each disc forms a capacitor C as shown in Fig. 8.10 (ii). This is known as mutual capacitance or self-capacitance. If there were mutual capacitance alone, then charging current would have been the same through all the discs and consequently voltage across each unit would have been the same i.e., V/3 as shown in Fig. 8.10 (ii). However, in actual practice, capacitance also exists between metal fitting of each disc and tower or earth. This is known as shunt capacitance C1. Due to shunt capacitance, charging current is not the same through all the discs of the string [See Fig. 8.10 (iii)]. Therefore, voltage across each disc will be different. Obviously, the disc nearest to the line conductor will have the maximum* voltage. Thus referring to Fig. 8.10 (iii), V3 will be much more than V2 or V1.

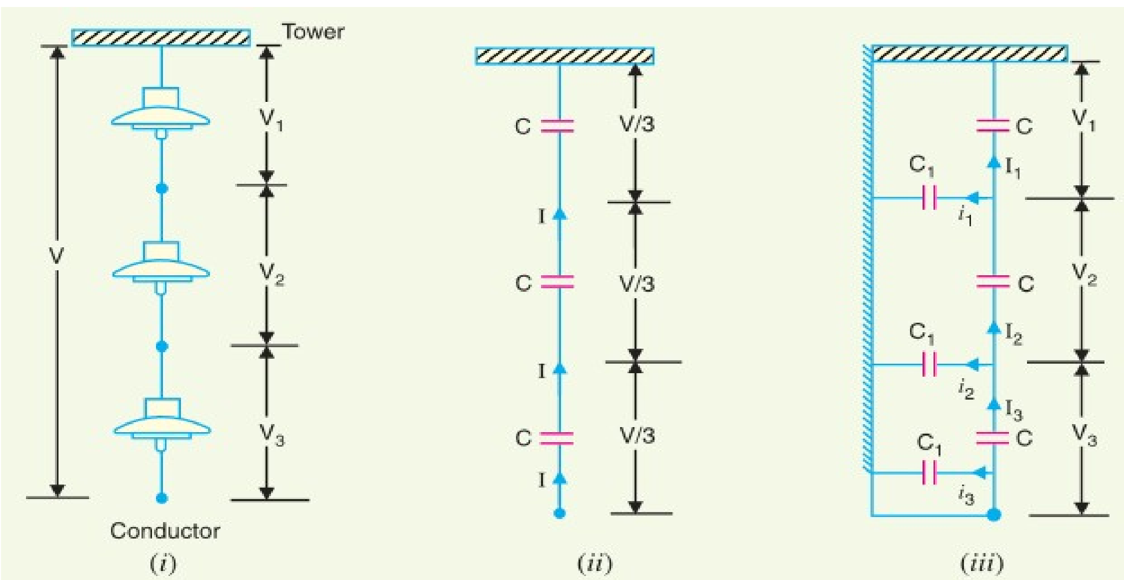


Fig. 8.10

The following points may be noted regarding the potential distribution over a string ofsuspension insulators :

(i) The voltage impressed on a string of suspension insulators does not distribute itself uniformly across the individual discs due to the presence of shunt capacitance.

(ii) The disc nearest to the conductor has maximum voltage across it. As we move towards the cross-arm, the voltage across each disc goes on decreasing.

(iii) The unit nearest to the conductor is under maximum electrical stress and is likely to be punctured. Therefore, means must be provided to equalize the potential across each unit.

(iv) If the voltage impressed across the string were d.c., then voltage across each unit would be the same. It is because insulator capacitances are ineffective for d.c.

String Efficiency:-

The ratio of voltage across the whole string to the product of number of discs and the voltage across the disc nearest to the conductor is known as string efficiency i.e.,

$$\text{String efficiency} = \frac{\text{Voltage across the string}}{n \times \text{Voltage across disc nearest to conductor}}$$

where n = number of discs in the string.

The greater the string efficiency, the more uniform is the voltage distribution.

Methods Of Improving String Efficiency:-

It has been seen above that potential distribution in a string of suspension insulators is not uniform. The maximum voltage appears across the insulator nearest to the line conductor and decreases progressively as the cross arm is approached. If the insulation of the highest stressed insulator (i.e. nearest to conductor) breaks down or flash over takes place.

So to equalize the potential across the various units of the string i.e. to improve the string efficiency. The various methods for this purpose are:

(i) By using longer cross-arms. The value of string efficiency depends upon the value of K i.e., ratio of shunt capacitance to mutual capacitance. The lesser the value of K, the greater is the string efficiency and more uniform is the voltage distribution. The value of K can be decreased by reducing the shunt capacitance. In order to reduce shunt capacitance, the distance of conductor from tower must be increased i.e., longer cross-arms should be used.

(ii) By grading the insulators. In this method, insulators of different dimensions are so chosen that each has a different capacitance. The insulators are capacitance graded i.e. they are assembled in the string in such a way that the top unit has the minimum capacitance, increasing progressively as the bottom unit (i.e., nearest to conductor) is reached. Since voltage is inversely proportional to capacitance, this method tends to equalize the potential distribution across the units in the string. This method has the disadvantage that a large number of different-sized insulators are required.

(iii) By using a guard ring. The potential across each unit in a string can be equalized by using a guard ring which is a metal ring electrically connected to the conductor and surrounding the bottom insulator as shown in the Fig. 8.13. The guard ring introduces capacitance between metal fittings and the line conductor. The guard ring is contoured in such a way that shunt capacitance currents i1, i2 etc. are equal to metal fitting line capacitance currents i’1, i’2 etc. The result is that same charging current I flows through each unit of string. Consequently, there will be uniform potential distribution across the units.

Derive for sag in overhead line with support at same level and different level (approximate formula effect of wind, ice and temperature on sag simple problem)

While erecting an overhead line, it is very important that conductors are under safe tension. In order to permit safe tension in the conductors, they are not fully stretched but are allowed to have a dip or sag.

"The difference in level between points of supports and the lowest point on the conductor is called sag."

Fig. 8.23. (i) Shows a conductor suspended between two equilevel supports A and B. The conductor is not fully stretched but is allowed to have a dip. The lowest point on the conductor is O and the sag is S. The following points may be noted :

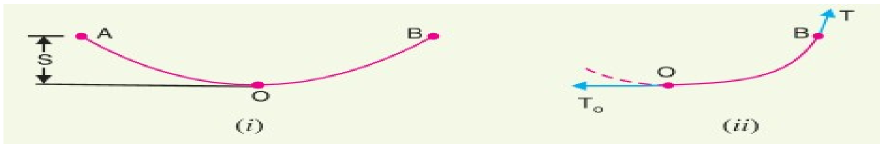


Fig. 8.23

(i) **When the conductor is suspended between two supports at the same level, it takes the shapeof catenary. However, if the sag is very small compared with the span, then sag-span curveis like a parabola.**

(ii) **The tension at any point on the conductor acts tangentially. Thus tension T_O at the lowest point O acts horizontally as shown in Fig. 8.23. (ii).**

(iii) **The horizontal component of tension is constant throughout the length of the wire.**

(iv) **The tension at supports is approximately equal to the horizontal tension acting at any point on the wire. Thus if T is the tension at the support B, then T = T_O.**

Calculation Of Sag

In an overhead line, the sag should be so adjusted that tension in the conductors is within safe limits. The tension is governed by conductor weight, effects of wind, ice loading and temperature variations. It is a standard practice to keep conductor tension less than 50% of its ultimate tensile strength i.e., minimum factor of safety in respect of conductor tension should be 2. We shall now calculate sag and tension of a conductor when (i) supports are at equal levels and (ii) supports are at unequal levels.

When supports are at equal levels.

Consider a conductor between two equilevel supports A and B with O as the lowest point as shown in Fig. 8.24. It can be proved that lowest point will be at the mid-span.

Let l= Length of span.
w= weight per unit length of conductor
T= Tension in the conductor

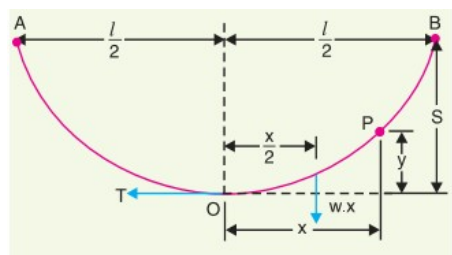


Fig. 8.24

Consider a point P on the conductor. Taking the lowest point O as the origin, let the co-ordinates of point P be x and y. Assuming that the curvature is so small that curved length is equal to its horizontal projection (i.e., OP = x), the two forces acting on the portion OP of the conductor are :

(a) **The weight wx of conductor acting at a distance x/2 from O.**

(b) **The tension T acting at O.**

Equating the moments of above two forces about point O, and we get,

$$Ty = wx \times \frac{x}{2}$$

or $$y = \frac{wx^2}{2T}$$

The maximum dip (sag) is represented by the value of y at either of the supports A and B.

At support A, $x = l/2$ and $y = S$

$\therefore$ $$\text{Sag, } S = \frac{w(l/2)^2}{2T} = \frac{wl^2}{8T}$$

When supports are at unequal levels.

In hilly areas, we generally come across conductors suspended between supports at unequal levels. Fig. 8.25 shows a conductor suspended between two supports A and B which are at different levels. The lowest point on the conductor is O.

Let

l = Span length

h = Difference in levels between two supports

x1 = Distance of support at lower level (i.e., A) from Ox2

= Distance of support at higher level (i.e. B) from OT =

Tension in the conductor

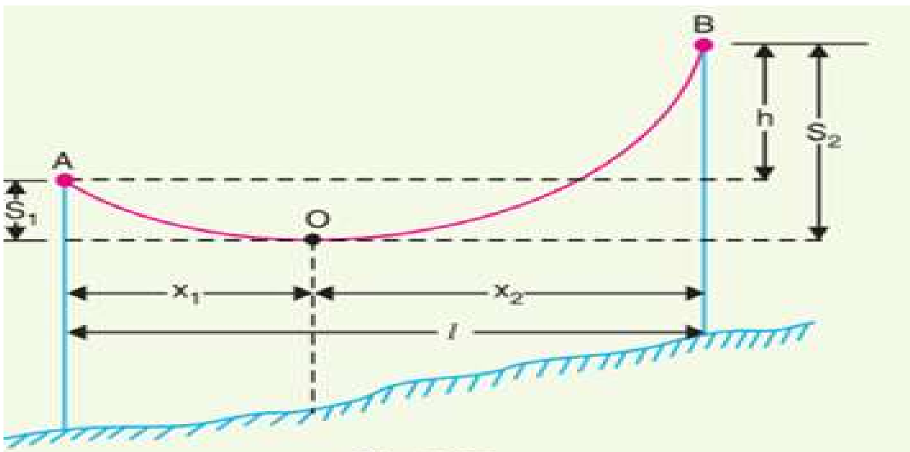


Fig. 8.25

If w is the weight per unit length of the conductor, then, Effect of wind and ice loading.

$$\text{Sag } S_1 = \frac{w\,x_1^{2*}}{2T}$$

and
$$\text{Sag } S_2 = \frac{w\,x_2^{2}}{2T}$$

Now
$$S_2 - S_1 = \frac{w}{2T}[x_2^2 - x_1^2] = \frac{w}{2T}(x_2 + x_1)(x_2 - x_1)$$

$$\therefore \quad S_2 - S_1 = \frac{w\,l}{2T}(x_2 - x_1) \qquad [\because x_1 + x_2 = l]$$

But
$$S_2 - S_1 = h$$

$$\therefore \quad h = \frac{w\,l}{2T}(x_2 - x_1)$$

or
$$x_2 - x_1 = \frac{2\,Th}{w\,l} \qquad \text{...}(ii)$$

Solving exps. (*i*) and (*ii*), we get,

$$x_1 = \frac{l}{2} - \frac{T\,h}{w\,l}$$

$$x_2 = \frac{l}{2} + \frac{T\,h}{w\,l}$$

Having found x_1 and x_2, values of S_1 and S_2 can be easily calculated.

The above formulae for sag are true only in still air and at normal temperature when the conductor is acted by its weight only. However, in actual practice, a conductor may have ice coating and simultaneously subjected to wind pressure. The weight of ice acts vertically downwards i.e., in the same direction as the weight of conductor. The force due to the wind is assumed to act horizontally i.e., at right angle to the projected surface of the conductor. Hence, the total force on the conductor is the vector sum of horizontal and vertical forces as shown in Fig. 8.26 (iii).

Total weight of conductor per unit length is

$$w_t = \sqrt{(w + w_i)^2 + (w_w)^2}$$

where

w = weight of conductor per unit length
= conductor material density × volume per unit length

w_i = weight of ice per unit length
= density of ice × volume of ice per unit length
= density of ice × $\frac{\pi}{4}[(d+2t)^2 - d^2] \times 1$
= density of ice × $\pi t(d+t)$*

w_w = wind force per unit length
= wind pressure per unit area × projected area per unit length
= wind pressure × $[(d+2t) \times 1]$

When the conductor has wind and ice loading also, the following points may be noted :

(i) The conductor sets itself in a plane at an angle θ to the vertical where

$$\tan\theta = \frac{w_w}{w + w_i}$$

(ii) The sag in the conductor is given by :

$$S = \frac{w_t l^2}{2T}$$

Hence S represents the slant sag in a direction making an angle θ to the vertical. *If no specific mention is made in the problem, then slant slag is calculated by using the above formula.*

(iii) The vertical sag $= S\cos\theta$

Example 8.17. *A 132 kV transmission line has the following data :*

Wt. of conductor = 680 kg/km ; Length of span = 260 m

Ultimate strength = 3100 kg ; Safety factor = 2

Calculate the height above ground at which the conductor should be supported. Ground clearance required is 10 metres.

Solution.

Wt. of conductor/metre run, $w = 680/1000 = 0{\cdot}68$ kg

Working tension, $T = \frac{\text{Ultimate strength}}{\text{Safety factor}} = \frac{3100}{2} = 1550$ kg

Span length, $l = 260$ m

$$\therefore \quad \text{Sag} = \frac{w l^2}{8T} = \frac{0{\cdot}68 \times (260)^2}{8 \times 1550} = 3{\cdot}7 \text{ m}$$

∴ Conductor should be supported at a height of 10 + 3·7 = **13·7 m**

Solution.

Span length, $l = 150$ m; Working tension, $T = 2000$ kg

Wind force/m length of conductor, $w_w = 1\cdot5$ kg

Wt. of conductor/m length, w = Sp. Gravity × Volume of 1 m conductor

$= 9\cdot9 \times 2 \times 100 = 1980$ gm $= 1\cdot98$ kg

Total weight of 1 m length of conductor is

$$w_t = \sqrt{w^2 + w_w^2} = \sqrt{(1\cdot98)^2 + (1\cdot5)^2} = 2\cdot48 \text{ kg}$$

$$\therefore \quad \text{Sag,} \quad S = \frac{w_t l^2}{8T} = \frac{2\cdot48 \times (150)^2}{8 \times 2000} = \mathbf{3\cdot48 \text{ m}}$$

This is the value of slant sag in a direction making an angle θ with the vertical. Referring to Fig. 8.27, the value of θ is given by ;

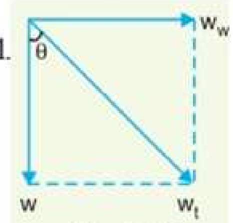


Fig. 8.27

$$\tan\theta = w_w/w = 1\cdot5/1\cdot98 = 0\cdot76$$

$$\therefore \quad \theta = \tan^{-1} 0\cdot76 = 37\cdot23°$$

$$\therefore \quad \text{Vertical sag} = S\cos\theta$$

$$= 3\cdot48 \times \cos 37\cdot23° = \mathbf{2\cdot77 \text{ m}}$$

Example 8.19. *A transmission line has a span of 200 metres between level supports. The conductor has a cross-sectional area of 1·29 cm², weighs 1170 kg/km and has a breaking stress of 4218 kg/cm². Calculate the sag for a safety factor of 5, allowing a wind pressure of 122 kg per square metre of projected area. What is the vertical sag?*

Solution.

Span length, $l = 200$ m

Wt. of conductor/m length, $w = 1170/1000 = 1\cdot17$ kg

Working tension, $*T = 4218 \times 1\cdot29/5 = 1088$ kg

Diameter of conductor, $d = \sqrt{\dfrac{4 \times \text{area}}{\pi}} = \sqrt{\dfrac{4 \times 1\cdot29}{\pi}} = 1\cdot28$ cm

Wind force/m length, w_w = Pressure × projected area in m²

$= (122) \times (1\cdot28 \times 10^{-2} \times 1) = 1\cdot56$ kg

Total weight of conductor per metre length is

$$w_t = \sqrt{w^2 + w_w^2} = \sqrt{(1\cdot17)^2 + (1\cdot56)^2} = 1\cdot95 \text{ kg}$$

$$\therefore \quad \text{Slant sag,} \quad S = \frac{w_t l^2}{8T} = \frac{1\cdot95 \times (200)^2}{8 \times 1088} = \mathbf{8\cdot96 \text{ m}}$$

The slant sag makes an angle θ with the vertical where value of θ is given by :

$$\theta = \tan^{-1}(w_w/w) = \tan^{-1}(1\cdot56/1\cdot17) = 53\cdot13°$$

$$\therefore \quad \text{Vertical sag} = S\cos\theta = 8\cdot96 \times \cos 53\cdot13° = \mathbf{5\cdot37 \text{ m}}$$

Example 8.20. *A transmission line has a span of 275 m between level supports. The conductor has an effective diameter of 1·96 cm and weighs 0·865 kg/m. Its ultimate strength is 8060 kg. If the conductor has ice coating of radial thickness 1·27 cm and is subjected to a wind pressure of 3·9 gm/cm² of projected area, calculate sag for a safety factor of 2. Weight of 1 c.c. of ice is 0·91 gm.*

Solution.

Span length, $l = 275$ m ; Wt. of conductor/m length, $w = 0{\cdot}865$ kg

Conductor diameter, $d = 1{\cdot}96$ cm ; Ice coating thickness, $t = 1{\cdot}27$ cm

Working tension, $T = 8060/2 = 4030$ kg

Volume of ice per metre (*i.e.*, 100 cm) length of conductor

$$= \pi t (d + t) \times 100 \text{ cm}^3$$

$$= \pi \times 1{\cdot}27 \times (1{\cdot}96 + 1{\cdot}27) \times 100 = 1288 \text{ cm}^3$$

Weight of ice per metre length of conductor is

$$w_i = 0{\cdot}91 \times 1288 = 1172 \text{ gm} = 1{\cdot}172 \text{ kg}$$

Wind force/m length of conductor is

$$w_w = [\text{Pressure}] \times [(d + 2t) \times 100]$$

$$= [3{\cdot}9] \times (1{\cdot}96 + 2 \times 1{\cdot}27) \times 100 \text{ gm} = 1755 \text{ gm} = 1{\cdot}755 \text{ kg}$$

Total weight of conductor per metre length of conductor is

$$w_t = \sqrt{(w + w_i)^2 + (w_w)^2}$$

$$= \sqrt{(0{\cdot}865 + 1{\cdot}172)^2 + (1{\cdot}755)^2} = 2{\cdot}688 \text{ kg}$$

$$\therefore \quad \text{Sag} = \frac{w_t l^2}{8T} = \frac{2{\cdot}688 \times (275)^2}{8 \times 4030} = \mathbf{6{\cdot}3 \text{ m}}$$

Example 8.22. *An overhead line has a span of 150 m between level supports. The conductor has a cross-sectional area of 2 cm². The ultimate strength is 5000 kg/cm² and safety factor is 5. The specific gravity of the material is 8·9 gm/cc. The wind pressure is 1·5 kg/m. Calculate the height of the conductor above the ground level at which it should be supported if a minimum clearance of 7 m is to be left between the ground and the conductor.*

Solution.

Span length, $l = 150$ m ; Wind force/m run, $w_w = 1{\cdot}5$ kg

Wt. of conductor/m run, w = conductor area × 100 cm × sp. gravity

$= 2 \times 100 \times 8{\cdot}9 = 1780 \text{ gm} = 1{\cdot}78$ kg

Working tension, $T = 5000 \times 2/5 = 2000$ kg

Total weight of one metre length of conductor is

$$w_t = \sqrt{w^2 + w_w^2} = \sqrt{(1{\cdot}78)^2 + (1{\cdot}5)^2} = 2{\cdot}33 \text{ kg}$$

$$\text{Slant sag, } S = \frac{w_t l^2}{8T} = \frac{2{\cdot}33 \times (150)^2}{8 \times 2000} = 3{\cdot}28 \text{ m}$$

$$\text{Vertical sag} = S \cos\theta = 3{\cdot}28 \times w/w_t = 3{\cdot}28 \times 1{\cdot}78/2{\cdot}33 = 2{\cdot}5 \text{ m}$$

Conductor should be supported at a height of 7 + 2·5 = **9·5 m**

Example 8.23. *The towers of height 30 m and 90 m respectively support a transmission line conductor at water crossing. The horizontal distance betwen the towers is 500 m. If the tension in the conductor is 1600 kg, find the minimum clearance of the conductor and water and clearance mid-way between the supports. Weight of conductor is 1·5 kg/m. Bases of the towers can be considered to be at water level.*

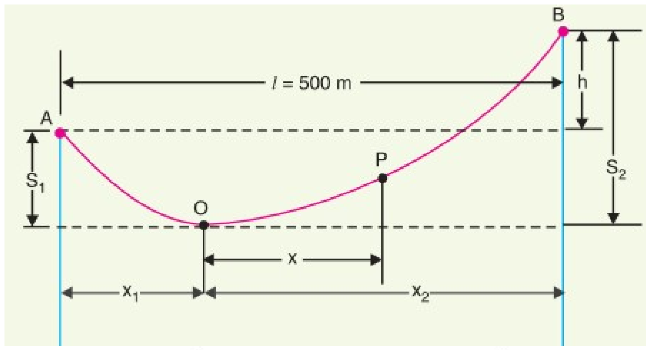
Solution.

Fig. 8.28 shows the conductor suspended between two supports A and B at different levels with O as the lowest point on the conductor.

Here, l = 500 m, w = 1·5 kg , T = 1600 kg. Difference in levels between supports,h = 90 − 30 = 60 m.

Let the lowest point O of the conductor be at a distance x1 from the support at lower level (i.e., support A) and at a distance x2 from the support at higher level (i.e., support B).

Obviously, x1 + x2 = 500 m ...(i)



Now Sag $S_1 = \frac{w x_1^2}{2T}$ and Sag $S_2 = \frac{w x_2^2}{2T}$

$$\therefore \quad h = S_2 - S_1 = \frac{w x_2^2}{2T} - \frac{w x_1^2}{2T}$$

or
$$60 = \frac{w}{2T}(x_2 + x_1)(x_2 - x_1)$$

$$\therefore \quad x_2 - x_1 = \frac{60 \times 2 \times 1600}{1 \cdot 5 \times 500} = 256 \text{ m} \qquad ...(ii)$$

Solving exps. (*i*) and (*ii*), we get, $x_1 = 122$ m; $x_2 = 378$ m

Now, $$S_1 = \frac{w x_1^2}{2T} = \frac{1 \cdot 5 \times (122)^2}{2 \times 1600} = 7 \text{ m}$$

Clearance of the lowest point *O* from water level

$$= 30 - 7 = \mathbf{23\ m}$$

Let the mid-point *P* be at a distance *x* from the lowest point *O*.

Clearly, $$x = 250 - x_1 = 250 - 122 = 128 \text{ m}$$

Sag at mid-point *P*, $$S_{mid} = \frac{w x^2}{2T} = \frac{1 \cdot 5 \times (128)^2}{2 \times 1600} = 7 \cdot 68 \text{ m}$$

Clearance of mid-point P from water level = 23 + 7·68 = 30·68 m (Ans)

Example 8.24. *An overhead transmission line conductor having a parabolic configuration weighs 1·925 kg per metre of length. The area of X-section of the conductor is 2·2 cm² and the ultimate strength is 8000 kg/cm². The supports are 600 m apart having 15 m difference of levels. Calculate the sag from the taller of the two supports which must be allowed so that the factor of safety shall be 5. Assume that ice load is 1 kg per metre run and there is no wind pressure.*

Solution. Fig. 8.29. shows the conductor suspended between two supports at A and B at different levels with O as the lowest point on the conductor.

Here, $l = 600$ m ; $w_i = 1$ kg ; $h = 15$m

$w = 1{\cdot}925$ kg ; $T = 8000 \times 2{\cdot}2/5 = 3520$ kg

Total weight of 1 m length of conductor is

$$w_t = w + w_i = 1{\cdot}925 + 1 = 2{\cdot}925 \text{ kg}$$

Let the lowest point O of the conductor be at a distance x_1 from the support at lower level (*i.e., A*) and at a distance x_2 from the support at higher level (*i.e., B*).

Clearly, $$x_1 + x_2 = 600 \text{ m} \quad ...(i)$$

Now, $$h = S_2 - S_1 = \frac{w_t\, x_2^2}{2T} - \frac{w_t\, x_1^2}{2T}$$

or $$15 = \frac{w_t}{2T}(x_2 + x_1)(x_2 - x_1)$$

$$\therefore \quad x_2 - x_1 = \frac{2 \times 15 \times 3520}{2{\cdot}925 \times 600} = 60 \text{ m} \quad ...(ii)$$

Solving exps. (*i*) and (*ii*), we have, $x_1 = 270$ m and $x_2 = 330$ m

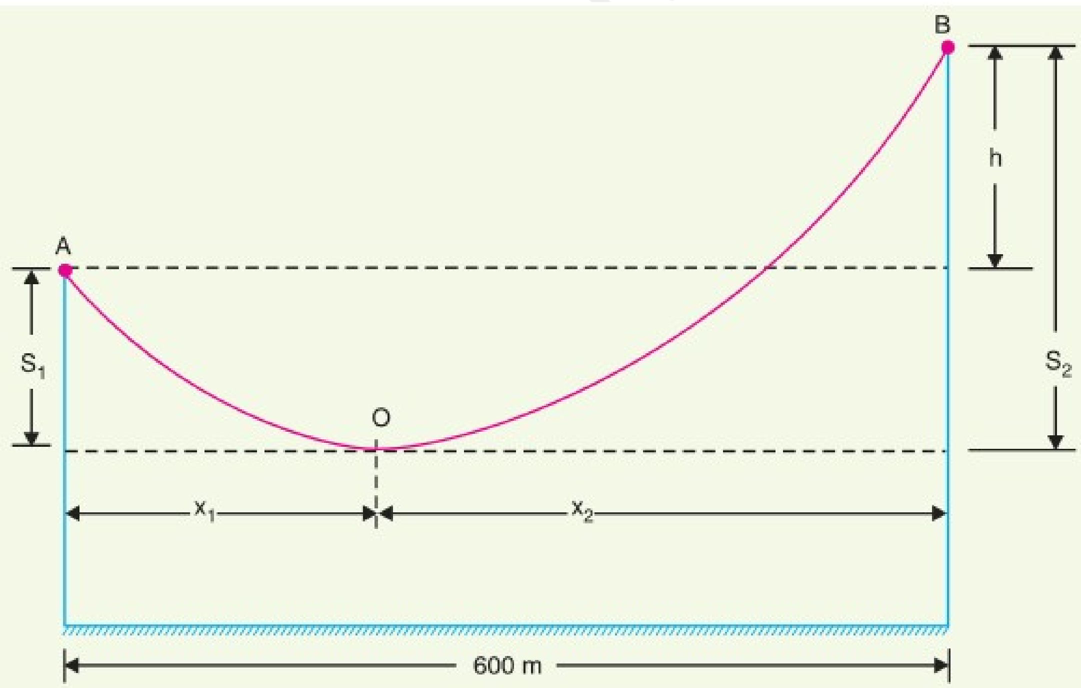


Fig. 8.29

Sag from the taller of the two towers is

$$S_2 = \frac{w_t\, x_2^2}{2T} = \frac{2{\cdot}925 \times (330)^2}{2 \times 3520} = \mathbf{45{\cdot}24 \text{ m}}$$

CHAPTER-4
PERFORMANCE OF SHORT & MEDIUM LINES

Calculation of regulation and efficiency.

The important considerations in the design and operation of a transmission line are the determination of voltage drop, line losses and efficiency of transmission. These values are greatly influenced by the line constants R, L and C of the transmission line.

CLASSIFICATION OF OVERHEAD TRANSMISSION LINES

Depending upon the manner in which capacitance is taken into account; the overhead transmission lines are classified as:

(i) Short transmission lines: When the length of an overhead transmission line is up to about 50 km and the line voltage is comparatively low (< 20 kV), it is usually considered as a short transmission line. Due to smaller length and lower voltage, the capacitance effects are small and hence can be neglected.

(ii) Medium transmission lines: When the length of an overhead transmission line is about 50km to 150 km and the line voltage is moderately high (>20 kV < 100 kV), it is considered as a medium transmission line. Due to sufficient length and voltage of the line, the capacitance effects are taken into account.

(iii) Long transmission lines: When the length of an overhead transmission line is more than 150 km and line voltage is very high (> 100 kV), it is considered as a long transmission line.

IMPORTANT TERMS:-

During the performance of a transmission line, it is desirable to determine its voltage regulation and transmission efficiency.

Voltage regulation: When a transmission line is carrying current, there is a voltage drop in the line due to resistance and inductance of the line. The result is that receiving end voltage (V_R) of the line is generally less than the sending end voltage (V_S).

"The difference in voltage at the receiving end of a transmission line between conditions of no load and full load is called voltage regulation and is expressed as a percentage of the receiving end voltage."

Transmission efficiency: The power obtained at the receiving end of a transmission line is generally less than the sending end power due to losses in the line resistance.

"The ratio of receiving end power to the sending end power of a transmission line is known as the transmission efficiency of the line" i.e.

$$\% \text{age Transmission efficiency, } \eta_T = \frac{\text{Receiving end power}}{\text{Sending end power}} \times 100$$

$$= \frac{V_R I_R \cos \varphi_R}{V_S I_S \cos \varphi_S} \times 100$$

where V_R, I_R and cos φ_R are the receiving end voltage, current and power factor while V_S, I_S and cos φ_S are the corresponding values at the sending end.

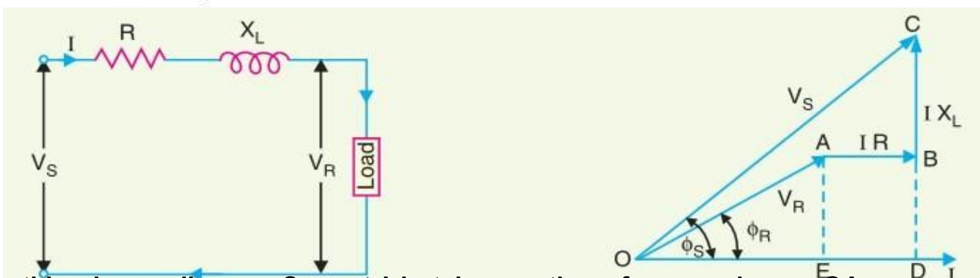
PERFORMANCE OF SINGLE PHASE SHORT TRANSMISSION LINE:

The effects of line capacitance are neglected for a short transmission line. Therefore, while studying the performance of such a line, only resistance and inductance of the line are taken into account. The equivalent circuit of a single phase short transmission line is shown in Fig.

Here, the total line resistance and inductance are shown as concentrated or lumped instead of being distributed. The circuit is a simple A.C series circuit.\

Let

$$
\begin{aligned}
I &= \text{load current} \\
R &= \text{loop resistance } i.e., \text{ resistance of both conductors} \\
X_L &= \text{loop reactance} \\
V_R &= \text{receiving end voltage} \\
\cos \varphi_R &= \text{receiving end power factor (lagging)} \\
V_S &= \text{sending end voltage}
\end{aligned}
$$



In this phasor diagram Current I is taken as the reference phasor. OA represents the receiving end voltage V_R leading I by φ_R. AB represents the drop IR in phase with I. BC represents the inductive drop IX_L and leads I by 90°. OC represents the sending end voltage V_S and leads I by φ_S. From the right angled triangle ODC, And we get,

$$
\begin{aligned}
(OC)^2 &= (OD)^2 + (DC)^2 \\
\text{or} \quad V_S^2 &= (OE + ED)^2 + (DB + BC)^2 \\
&= (V_R \cos \varphi_R + IR)^2 + (V_R \sin \varphi_R + IX_L)^2 \\
\therefore \quad V_S &= \sqrt{(V_R \cos \varphi_R + IR)^2 + (V_R \sin \varphi_R + IX_L)^2}
\end{aligned}
$$

$$(i) \quad \text{\%age Voltage regulation} = \frac{V_S - V_R}{V_R} \times 100$$

$$(ii) \quad \text{Sending end } p.f., \cos \varphi_S = \frac{OD}{OC} = \frac{V_R \cos \varphi_R + IR}{V_S}$$

$$(iii) \quad \text{Power delivered} = V_R I_R \cos \varphi_R$$

$$\text{Line losses} = I^2 R$$

$$\text{Power sent out} = V_R I_R \cos \varphi_R + I^2 R$$

$$\text{\%age Transmission efficiency} = \frac{\text{Power delivered}}{\text{Power sent out}} \times 100$$

$$= \frac{V_R I_R \cos \varphi_R}{V_R I_R \cos \varphi_R + I^2 R} \times 100$$

An approximate expression for the sending end voltage VS can be obtained as follows. Draw perpendicular from B and C on OA produced as shown in Fig. Then OC is nearly equal to OF i.e.

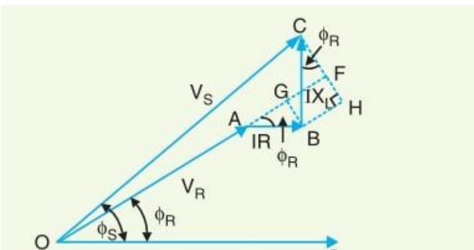


Fig. 10.2

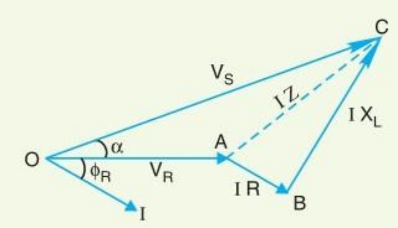


Fig. 10.3

$$OC = OF = OA + AF = OA + AG + GF$$

$$= OA + AG + BH$$

$$V_S = V_R + IR \cos \varphi_R + IX_L \sin \varphi_R$$

<u>Solution in complex notation:</u> It is often convenient and profitable to make the line calculations in complex notation.

Taking $\vec{V_R}$ as the reference phasor, draw the phasor diagram as shown in Fig 10.3. It is clear that $\vec{V_S}$ is the phasor sum of $\vec{V_R}$ and $\vec{I}\,\vec{Z}$.

$$*\vec{V_R} = V_R + j\,0$$

$$\vec{I} = \vec{I} \angle -\varphi_R = I(\cos \varphi_R - j \sin \varphi_R)$$

$$\vec{Z} = R + jX_L$$

$$\therefore \quad \vec{V_S} = \vec{V_R} + \vec{I}\,\vec{Z}$$

$$= (V_R + j\,0) + I(\cos \varphi_R - j \sin \varphi_R)(R + jX_L)$$

$$= (V_R + IR \cos \varphi_R + IX_L \sin \varphi_R) + j(IX_L \cos \varphi_R - IR \sin \varphi_R)$$

$$\therefore \quad V_S = \sqrt{(V_R + IR \cos \varphi_R + IX_L \sin \varphi_R)^2 + (IX_L \cos \varphi_R - IR \sin \varphi_R)^2}$$

The second term under the root is quite small and can be neglected with reasonable accuracy. Therefore, approximate expression for V_S becomes :

$$V_S = V_R + IR \cos \varphi_R + IX_L \sin \varphi_R$$

The following points may be noted :

(i) The approximate formula for $V_S = (V_R + I\,R\cos\varphi_R + I\,X_L\sin\varphi_R)$ gives fairly correct results for lagging power factors. However, appreciable error is caused for leading power factors. Therefore, approximate expression for V_S should be used for lagging p.f. only.

(ii) The solution in complex notation is in more presentable form.

Effect of Load P.F. on Regulation & Efficiency:

The regulation and efficiency of a transmission line depend to a considerable extent upon the power factor of the load.

1. Effect on regulation. The expression for voltage regulation of a short transmission line is given by :

$$\text{\%age Voltage regulation} = \frac{I\,R\cos\varphi_R + I\,X_L\sin\varphi_R}{V_R} \times 100 \qquad (\textit{for lagging p.f.})$$

$$\text{\%age Voltage regulation} = \frac{I\,R\cos\varphi_R - I\,X_L\sin\varphi_R}{V_R} \times 100 \qquad (\textit{for leading p.f.})$$

The following conclusions can be drawn from the above expressions :

(*i*) When the load p.f. is lagging or unity or such leading that $I\,R\cos\varphi_R > I\,X_L\sin\varphi_R$, then voltage regulation is positive *i.e.*, receiving end voltage V_R will be less than the sending end voltage V_S.

(*ii*) For a given V_R and I, the voltage regulation of the line increases with the decrease in p.f. for lagging loads.

(*iii*) When the load p.f. is leading to this extent that $I\,X_L\sin\varphi_R > I\,R\cos\varphi_R$, then voltage regulation is negative *i.e.* the receiving end voltage V_R is more than the sending end voltage V_S.

(*iv*) For a given V_R and ... decrease in p.f. for leading loads.

2. Effect on transm ... pends upon the power factor.

$$P = V_R * I\cos\varphi_R \quad \text{(For 1-phase line)}$$

$$\therefore \quad I = \frac{P}{V_R\cos\varphi_R}$$

$$P = 3\,V_R\,I\cos\varphi_R \quad \text{(For 3-phase line)}$$

$$I = \frac{P}{3\,V_R\cos\varphi_R}$$

It is clear that in each case, for a given amount of power to be transmitted (P) and receiving end voltage (VR), the load current I is inversely proportional to the load p.f. $\cos\varphi_R$. Consequently, with the decrease in load p.f., the load current and hence the line losses are increased.

This leads to the conclusion that transmission efficiency of a line decreases with the decrease in load p.f. and vice-versa.

Example 10.1. *A single phase overhead transmission line delivers 1100 kW at 33 kV at 0·8 p.f. lagging. The total resistance and inductive reactance of the line are 10 Ω and 15 Ω respectively. Determine : (i) sending end voltage (ii) sending end power factor and (iii) transmission efficiency.*

Solution.

Load power factor, $\cos \phi_R = 0{\cdot}8$ lagging

Total line impedance, $\vec{Z} = R + jX_L = 10 + j\,15$

Receiving end voltage, $V_R = 33$ kV $= 33{,}000$ V

$$\therefore \quad \text{Line current, } I = \frac{kW \times 10^3}{V_R \cos \phi_R} = \frac{1100 \times 10^3}{33{,}000 \times 0{\cdot}8} = 41{\cdot}67 \text{ A}$$

$$\text{As } \cos \phi_R = 0{\cdot}8 \quad \therefore \quad \sin \phi_R = 0{\cdot}6$$

The equivalent circuit & phasor diagram as shown in Fig. (i) & (ii) Taking receiving end voltage V_R as the reference phasor,

$$\vec{V_R} = V_R + j\,0 = 33000 \text{ V}$$

$$\vec{I} = I(\cos \phi_R - j \sin \phi_R)$$

$$= 41{\cdot}67\,(0{\cdot}8 - j\,0{\cdot}6) = 33{\cdot}33 - j\,25$$

(*i*) Sending end voltage, $\vec{V_S} = \vec{V_R} + \vec{I}\,\vec{Z}$

$$= 33{,}000 + (33{\cdot}33 - j\,25{\cdot}0)\,(10 + j\,15)$$

$$= 33{,}000 + 333{\cdot}3 - j250 + j500 + 375$$

$$= 33{,}708{\cdot}3 + j\,250$$

$\therefore$ Magnitude of $V_S = \sqrt{(33{,}708{\cdot}3)^2 + (250)^2} =$ **33,709 V**

(*ii*) Angle between $\vec{V_S}$ and $\vec{V_R}$ is

$$\alpha = \tan^{-1} \frac{250}{33{,}708{\cdot}3} = \tan^{-1} 0{\cdot}0074 = 0{\cdot}42°$$

$\therefore$ Sending end power factor angle is

$$\phi_S = \phi_R + \alpha = 36{\cdot}87° + 0{\cdot}42° = 37{\cdot}29°$$

$\therefore$ Sending end p.f., $\cos \phi_S = \cos 37{\cdot}29° =$ **0·7956 lagging**

(*iii*) Line losses $= I^2R = (41{\cdot}67)^2 \times 10 = 17{,}364$ W $= 17{\cdot}364$ kW

Output delivered $= 1100$ kW

Power sent $= 1100 + 17{\cdot}364 = 1117{\cdot}364$ kW

$\therefore$ Transmission efficiency $= \dfrac{\text{Power delivered}}{\text{Power sent}} \times 100 = \dfrac{1100}{1117{\cdot}364} \times 100 =$ **98·44%**

Note. V_S and ϕ_S can also be calculated as follows :

$$V_S = V_R + IR\cos\phi_R + IX_L \sin\phi_R \text{ (approximately)}$$

$$= 33{,}000 + 41{\cdot}67 \times 10 \times 0{\cdot}8 + 41{\cdot}67 \times 15 \times 0{\cdot}6$$

$$= 33{,}000 + 333{\cdot}36 + 375{\cdot}03$$

$$= 33708{\cdot}39 \text{ V which is approximately the same as above}$$

$$\cos\phi_S = \frac{V_R \cos\phi_R + IR}{V_S} = \frac{33{,}000 \times 0{\cdot}8 + 41{\cdot}67 \times 10}{33{,}708{\cdot}39} = \frac{26{,}816{\cdot}7}{33{,}708{\cdot}39}$$

$$= 0{\cdot}7958$$

Example 10.2. *What is the maximum length in km for a 1-phase transmission line having copper conductor of 0·775 cm² cross-section over which 200 kW at unity power factor and at 3300V are to be delivered ? The efficiecny of transmission is 90%. Take specific resistance as 1.725 μ Ω cm.*

Solution.

$$\text{Receiving end power} = 200 \text{ kW} = 2,00,000 \text{ W}$$

$$\text{Transmission efficiency} = 0 \cdot 9$$

$$\therefore \quad \text{Sending end power} = \frac{2,00,000}{0 \cdot 9} = 2,22,222 \text{ W}$$

$$\therefore \quad \text{Line losses} = 2,22,222 - 2,00,000 = 22,222 \text{ W}$$

$$\text{Line current, } I = \frac{200 \times 10^3}{3,300 \times 1} = 60 \cdot 6 \text{ A}$$

Let R Ω be the resistance of one conductor.

$$\text{Line losses} = 2 I^2 R$$

or

$$22,222 = 2 (60 \cdot 6)^2 \times R$$

$$\therefore \quad R = \frac{22,222}{2 \times (60 \cdot 6)^2} = 3 \cdot 025 \ \Omega$$

Now,

$$R = \rho \, l/a$$

$$\therefore \quad l = \frac{Ra}{\rho} = \frac{3 \cdot 025 \times 0 \cdot 775}{1 \cdot 725 \times 10^{-6}} = 1 \cdot 36 \times 10^6 \text{ cm} = \mathbf{13 \cdot 6 \text{ km}}$$

Example 10.3. *An overhead 3-phase transmission line delivers 5000 kW at 22 kV at 0·8 p.f. lagging. The resistance and reactance of each conductor is 4 Ω and 6 Ω respectively. Determine : (i) sending end voltage (ii) percentage regulation (iii) transmission efficiency.*

Solution.

$$\text{Load power factor, } \cos \phi_R = 0 \cdot 8 \text{ lagging}$$

$$\text{Receiving end voltage/phase, *} V_R = 22,000/\sqrt{3} = 12,700 \text{ V}$$

$$\text{Impedance/phase, } \quad \vec{Z} = 4 + j\,6$$

$$\text{Line current, } \quad I = \frac{5000 \times 10^3}{3 \times 12700 \times 0 \cdot 8} = 164 \text{ A}$$

$$\text{As } \cos \phi_R = 0 \cdot 8 \quad \therefore \sin \phi_R = 0 \cdot 6$$

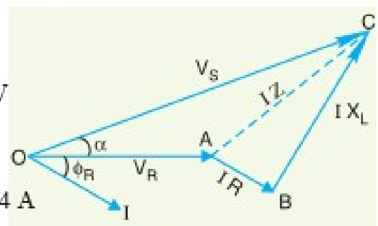


Fig. 10.6

Taking $\vec{V_R}$ as the reference phasor (see Fig. 10.6),

$$\vec{V_R} = V_R + j\,0 = 12700 \text{ V}$$

$$\vec{I} = I (\cos \phi_R - j \sin \phi_R) = 164 (0 \cdot 8 - j\,0 \cdot 6) = 131 \cdot 2 - j\,98 \cdot 4$$

(*i*) Sending end voltage per phase is

$$\begin{aligned} \vec{V_S} &= \vec{V_R} + \vec{I}\,\vec{Z} = 12700 + (131 \cdot 2 - j\,98 \cdot 4)(4 + j\,6) \\ &= 12700 + 524 \cdot 8 + j\,787 \cdot 2 - j\,393 \cdot 6 + 590 \cdot 4 \\ &= 13815 \cdot 2 + j\,393 \cdot 6 \end{aligned}$$

$$\text{Magnitude of } V_S = \sqrt{(13815 \cdot 2)^2 + (393 \cdot 6)^2} = 13820 \cdot 8 \text{ V}$$

$$\text{Line value of } V_S = \sqrt{3} \times 13820 \cdot 8 = 23938 \text{ V} = \mathbf{23 \cdot 938 \text{ kV}}$$

(*ii*)
$$\text{\% age Regulation} = \frac{V_S - V_R}{V_R} \times 100 = \frac{13820 \cdot 8 - 12700}{12700} \times 100 = \mathbf{8 \cdot 825\%}$$

(*iii*)
$$\text{Line losses} = 3 I^2 R = 3 \times (164)^2 \times 4 = 3,22,752 \text{ W} = 322 \cdot 752 \text{ kW}$$

$$\therefore \quad \text{Transmission efficiency} = \frac{5000}{5000 + 322 \cdot 752} \times 100 = \mathbf{93 \cdot 94\%}$$

PERFORMANCE OF MEDIUM TRANSMISSION LINE:

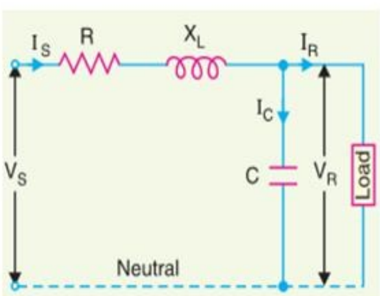
In short transmission line calculations, the effects of the line capacitance are neglected because such lines have smaller lengths and transmit power at relatively low voltages (< 20 kV). However, as the length and voltage of the line increase, the capacitance gradually becomes of greater importance. Since medium transmission lines have sufficient length (50-150 km) and usually operate at voltages greater than 20 kV, the effects of capacitance cannot be neglected.

The capacitance is uniformly distributed over the entire length of the line. However, in orderto make the calculations simple, the line capacitance is assumed to be lumped or concentrated in the form of capacitors shunted across the line at one or more points. The most commonly used methods (known as localized capacitance methods) for the solution of medium transmissions lines are:

(i) End Condenser Method.
(ii) Nominal T Method.
(iii) Nominal π Method.

End condenser method:

In this method, the capacitance of the line is lumped or concentrated at the receiving or load end as shown in Fig. This method of localizing the line capacitance at the load end overestimates the effects of capacitance.



Let I_R = load current per phase
R = resistance per phase
X_L = inductive reactance per phase
C = capacitance per phase
$\cos \varphi_R$ = receiving end power factor (*lagging*)

V_S = sending end voltage per phase

The *phasor diagram for the circuit is shown in Fig 10.9. Taking the receiving end voltage $\overrightarrow{V_R}$ as the reference phasor, we have, $\overrightarrow{V_R} = V_R + j\,0$

Load current, $\overrightarrow{I_R} = I_R\,(\cos \varphi_R - j \sin \varphi_R)$

Capacitive current, $\overrightarrow{I_C} = j\,\overrightarrow{V_R}\,\omega C = j\,2\,\pi f C\,\overrightarrow{V_R}$

The sending end current $\overrightarrow{I_S}$ is the phasor sum of load current $\overrightarrow{I_R}$ and capacitive current $\overrightarrow{I_C}$ *i.e.*,

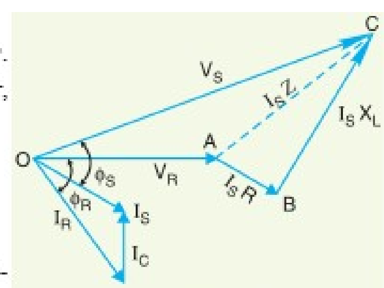


Fig. 10.9

$$\vec{I_S} = \vec{I_R} + \vec{I_C}$$
$$= I_R(\cos\varphi_R - j\sin\varphi_R) + j\,2\pi f C\,V_R$$
$$= I_R\cos\varphi_R + j(-I_R\sin\varphi_R + 2\pi f C V_R)$$

Voltage drop/phase $= \vec{I_S}\vec{Z} = \vec{I_S}(R + jX_L)$

Sending end voltage, $\vec{V_S} = \vec{V_R} + \vec{I_S}\vec{Z} = \vec{V_R} + \vec{I_S}(R + jX_L)$

Thus, the magnitude of sending end voltage V_S can be calculated.

$$\%\text{ Voltage regulation} = \frac{V_S - V_R}{V_R} \times 100$$

$$\%\text{ Voltage transmission efficiency} = \frac{\text{Power delivered / phase}}{\text{Power delivered / phase + losses / phase}} \times 100$$
$$= \frac{V_R I_R \cos\varphi_R}{V_R I_R \cos\varphi_R + I_S^2 R} \times 100$$

Limitations. Although end condenser method for the solution of medium lines is simple to work outcalculations, yet it has the following drawbacks :

(i) There is a considerable error (about 10%) in calculations because the distributed capacitance hasbeen assumed to be lumped or concentrated.

(ii) This method overestimates the effects of line capacitance.

Example 10.10. *A (medium) single phase transmission line 100 km long has the following constants :*

Resistance/km = 0·25 Ω ; *Reactance/km = 0·8 Ω*

Susceptance/km = 14 × 10^{-6} siemen ; *Receiving end line voltage = 66,000 V*

Assuming that the total capacitance of the line is localised at the receiving end alone, determine (i) the sending end current (ii) the sending end voltage (iii) regulation and (iv) supply power factor. The line is delivering 15,000 kW at 0.8 power factor lagging. Draw the phasor diagram to illustrate your calculations.

Total resistance, $R = 0{\cdot}25 \times 100 = 25\ \Omega$

Total reactance, $X_L = 0{\cdot}8 \times 100 = 80\ \Omega$

Total susceptance, $Y = 14 \times 10^{-6} \times 100 = 14 \times 10^{-4}\ S$

Receiving end voltage, $V_R = 66{,}000$ V

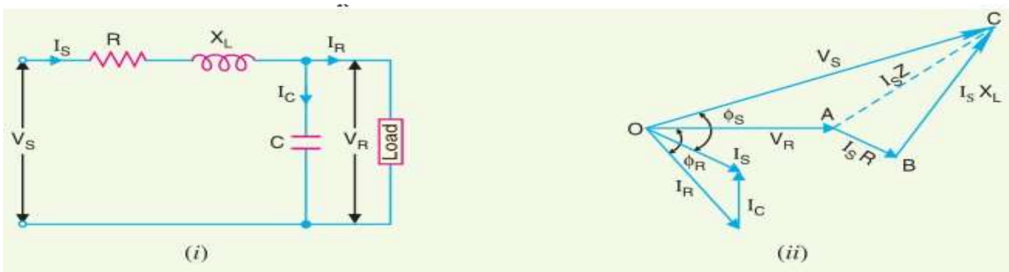
$\therefore$ Load current, $I_R = \dfrac{15{,}000 \times 10^3}{66{,}000 \times 0{\cdot}8} = 284$ A

$\cos\varphi_R = 0{\cdot}8$; $\sin\varphi_R = 0{\cdot}6$

Taking receiving end voltage as the reference phasor [see Fig.10.10 (*ii*)], we have,

$$\vec{V_R} = V_R + j\,0 = 66{,}000\,V$$

Load current, $\vec{I_R} = I_R(\cos\varphi_R - j\sin\varphi_R) = 284\,(0{\cdot}8 - j\,0{\cdot}6) = 227 - j\,170$



Capacitive current, $\qquad \vec{I_C} = j\,Y \times V_R = j\,14 \times 10^{-4} \times 66000 = j\,92$

(i) Sending end current, $\qquad \vec{I_S} = \vec{I_R} + \vec{I_C} = (227 - j\,170) + j\,92$

$$= 227 - j\,78 \qquad \text{... (i)}$$

Magnitude of $I_S = \sqrt{(227)^2 + (78)^2} =$ **240 A**

(ii) Voltage drop

$$= \vec{I_S}\,\vec{Z} = \vec{I_S}(R + j\,X_L) = (227 - j\,78)\,(25 + j\,80)$$
$$= 5{,}675 + j\,18,160 - j\,1950 + 6240$$
$$= 11{,}915 + j\,16{,}210$$

Sending end voltage, $\qquad \vec{V_S} = \vec{V_R} + \vec{I_S}\,\vec{Z} = 66{,}000 + 11{,}915 + j\,16{,}210$

$$= 77{,}915 + j\,16{,}210 \qquad \text{...(ii)}$$

Magnitude of $V_S = \sqrt{(77915)^2 + (16210)^2} =$ **79583V**

(iii) % Voltage regulation $= \dfrac{V_S - V_R}{V_R} \times 100 = \dfrac{79{,}583 - 66{,}000}{66{,}000} \times 100 =$ **20·58%**

(iv) Referring to exp. (i), phase angle between $\vec{V_R}$ and $\vec{I_R}$ is :

$$\theta_1 = \tan^{-1} - 78/227 = \tan^{-1}(-0{\cdot}3436) = -18{\cdot}96°$$

Referring to exp. (ii), phase angle between $\vec{V_R}$ and $\vec{V_S}$ is :

$$\theta_2 = \tan^{-1}\frac{16210}{77915} = \tan^{-1}(0{\cdot}2036) = 11{\cdot}50°$$

∴ Supply power factor angle, $\quad \varphi_S = 18{\cdot}96° + 11{\cdot}50° = 30{\cdot}46°$

∴ Supply p.f. $= \cos\varphi_S = \cos 30{\cdot}46° =$ **0·86 lag**

Nominal T Method:-

In this method, the whole line capacitance is assumed to be concentrated at the middle point of the line and half the line resistance and reactance are lumped on its either side as shown in Fig. 10.11. Therefore, in this arrangement, full charging current flows over half the line. In Fig. 10.11, one phaseof 3phase transmission line is shown as it is advantageous to work in phase instead of line-to-line values.

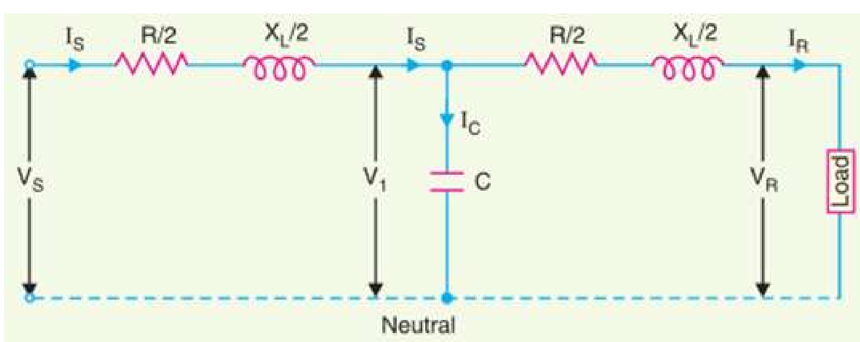


Fig. 10.11

Let $\quad I_R$ = load current per phase ; $\quad R$ = resistance per phase

X_L = inductive reactance per phase ; $\quad C$ = capacitance per phase

$\cos \phi_R$ = receiving end power factor (*lagging*) ; $\quad V_S$ = sending end voltage/phase

V_1 = voltage across capacitor C

The *phasor diagram for the circuit is shown in Fig. 10.12. Taking the receiving end voltage $\vec{V_R}$ as the reference phasor, we have,

Receiving end voltage, $\quad \vec{V_R} = V_R + j\,0$

Load current, $\quad \vec{I_R} = I_R(\cos \phi_R - j \sin \phi_R)$

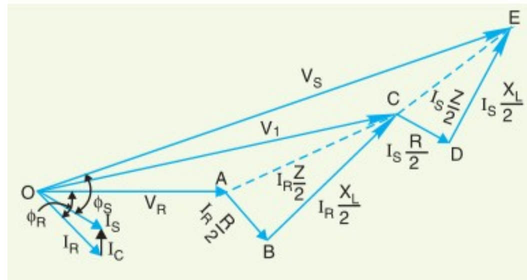


Fig. 10.12

Note the construction of phasor diagram. $\vec{V_R}$ is taken as the reference phasor represented by OA. The load current $\vec{I_R}$ lags behind $\vec{V_R}$ by ϕ_R. The drop $AB = I_R R/2$ is in phase with $\vec{I_R}$ and $BC = I_R X_L/2$ leads $\vec{I_R}$ by 90°. The phasor OC represents the voltage $\vec{V_1}$ across condenser C. The capacitor current $\vec{I_C}$ leads $\vec{V_1}$ by 90° as shown. The phasor sum of $\vec{I_R}$ and $\vec{I_C}$ gives $\vec{I_S}$. Now $CD = I_S R/2$ is in phase with $\vec{I_S}$ while $DE = I_S X_L/2$ leads $\vec{I_S}$ by 90°. Then, OE represents the sending end voltage $\vec{V_S}$.

Voltage across C, $\quad \vec{V_1} = \vec{V_R} + \vec{I_R}\,\vec{Z}/2$

$$= V_R + I_R(\cos \phi_R - j \sin \phi_R)\left(\frac{R}{2} + j\frac{X_L}{2}\right)$$

Capacitive current, $\quad \vec{I_C} = j\,\omega\,C\,\vec{V_1} = j\,2\pi f C\,\vec{V_1}$

Sending end current, $\quad \vec{I_S} = \vec{I_R} + \vec{I_C}$

Sending end voltage, $\quad \vec{V_S} = \vec{V_1} + \vec{I_S}\,\frac{\vec{Z}}{2} = \vec{V_1} + \vec{I_S}\left(\frac{R}{2} + j\frac{X_L}{2}\right)$

Example 10.11. *A 3-phase, 50-Hz overhead transmission line 100 km long has the following constants :*

Resistance/km/phase $= 0.1\ \Omega$

Inductive reactance/km/phase $= 0{\cdot}2\ \Omega$

Capacitive susceptance/km/phase $= 0{\cdot}04 \times 10^{-4}$ *siemen*

Determine (i) the sending end current (ii) sending end voltage (iii) sending end power factor and (iv) transmission efficiency when supplying a balanced load of 10,000 kW at 66 kV, p.f. 0·8 lagging. Use nominal T method.

Solution. Figs. 10.13 (*i*) and 10.13 (*ii*) show the circuit diagram and phasor diagram of the line respectively.

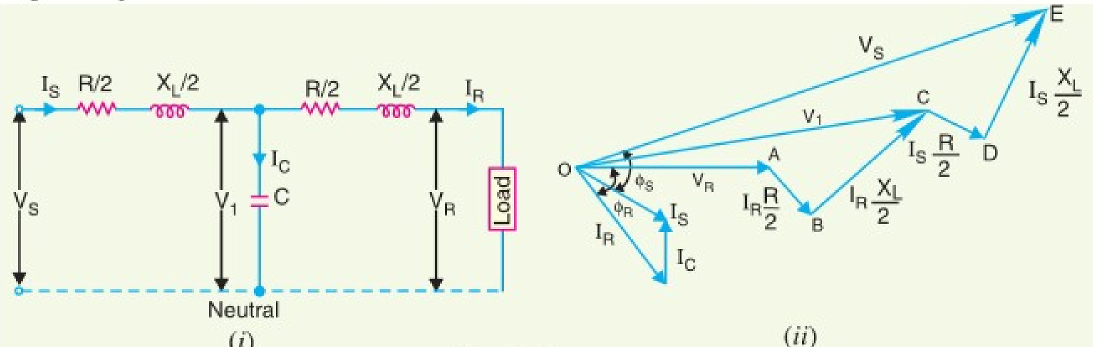


Fig. 10.13

Total resistance/phase, $R = 0{\cdot}1 \times 100 = 10\ \Omega$

Total reactance/phase. $X_L = 0{\cdot}2 \times 100 = 20\ \Omega$

Capacitive susceptance, $Y = 0{\cdot}04 \times 10^{-4} \times 100 = 4 \times 10^{-4}$ S

Receiving end voltage/phase, $V_R = 66{,}000/\sqrt{3} = 38105\ V$

Load current, $$I_R = \frac{10{,}000 \times 10^3}{\sqrt{3} \times 66 \times 10^3 \times 0 \cdot 8} = 109 \text{ A}$$

$$\cos \varphi_R = 0{\cdot}8\ ; \quad \sin \varphi_R = 0{\cdot}6$$

Impedance per phase, $\vec{Z} = R + j\,X_L = 10 + j\,20$

(*i*) Taking receiving end voltage as the reference phasor [see Fig. 10.13 (*ii*)], we have,

Receiving end voltage, $\vec{V_R} = V_R + j\,0 = 38{,}105\ V$

Load current, $\vec{I_R} = I_R\,(\cos \varphi_R - j \sin \varphi_R) = 109\,(0{\cdot}8 - j\,0{\cdot}6) = 87{\cdot}2 - j\,65{\cdot}4$

Voltage across C, $\vec{V_1} = \vec{V_R} + \vec{I_R}\,\vec{Z}/2 = 38,\,105 + (87{\cdot}2 \ \ -j\,65{\cdot}4)\,(5 + j\,10)$

$= 38{,}105 + 436 + j\,872 - j\,327 + 654 = 39{,}195 + j\,545$

Charging current, $\vec{I_C} = j\,Y\,\vec{V_1} = j\,4\times10^{-4}(39{,}195 + j\,545) = -0\cdot218 + j\,15\cdot6$

Sending end current, $\vec{I_S} = \vec{I_R} + \vec{I_C} = (87\cdot2 - j\,65\cdot4) + (-0\cdot218 + j\,15\cdot6)$

$= 87\cdot0 - j\,49\cdot8 = 100\,\angle\,-29°47'$ A

∴ Sending end current = **100 A**

(ii) Sending end voltage, $\vec{V_S} = \vec{V_1} + \vec{I_S}\,\vec{Z}/2 = (39{,}195 + j\,545) + (87\cdot0 - j\,49\cdot8)\,(5 + j\,10)$

$= 39{,}195 + j\,545 + 434\cdot9 + j\,870 - j\,249 + 498$

$= 40128 + j\,1170 = 40145\,\angle 1°40'$ V

∴ Line value of sending end voltage

$= 40145 \times \sqrt{3} = 69\,533$ V $=$ **69·533 kV**

(iii) Referring to phasor diagram in Fig. 10.14,

θ_1 = angle between $\vec{V_R}$ and $\vec{V_S}$ = 1°40′

θ_2 = angle between $\vec{V_R}$ and $\vec{I_S}$ = 29° 47′

∴ ϕ_S = angle between $\vec{V_S}$ and $\vec{I_S}$

$= \theta_1 + \theta_2 = 1°40' + 29°47' = 31°27'$

∴ Sending end power factor, $\cos\phi_S = \cos 31°27' =$ **0·853 lag**

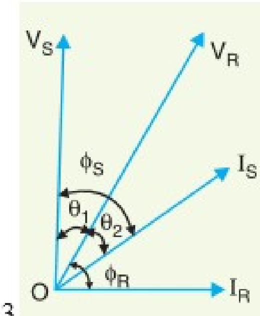


Fig. 10.14

(iv) Sending end power $= 3\,V_S I_S \cos\phi_S = 3\times40{,}145\times100\times0\cdot853$

$= 10273105$ W $= 10273\cdot105$ kW

Power delivered = 10,000 kW

∴ Transmission efficiency $= \dfrac{10{,}000}{10273\cdot105}\times100 =$ **97·34%**

Nominal π Method:-

In this method, capacitance of each conductor (i.e., line to neutral) is divided into two halves; one half being lumped at the sending end and the other half at the receiving end as shown in Fig. 10.16.It is obvious that capacitance at the sending end has no effect on the line drop. However, its charging current must be added to line current in order to obtain the total sending end current.

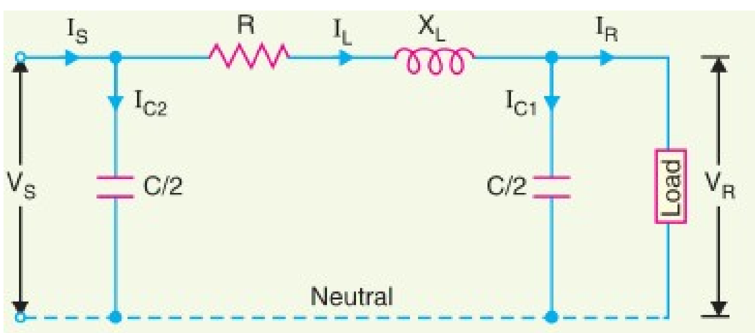


Fig. 10.16

Let $\qquad I_R$ = load current per phase

R = resistance per phase

X_L = inductive reactance per phase

C = capacitance per phase

$\cos \varphi_R$ = receiving end power factor (*lagging*)

V_S = sending end voltage per phase

The *phasor diagram for the circuit is shown in Fig. 10.17. Taking the receiving end voltage as the reference phasor, we have,

$$\overrightarrow{V_R} = V_R + j\,0$$

Load current, $\qquad \overrightarrow{I_R} = I_R(\cos \varphi_R - j \sin \varphi_R)$

Charging current at load end is

$$\overrightarrow{I_{C1}} = j\,\omega\,(C/2)\,\overrightarrow{V_R} = j\,\pi\,f\,C\,\overrightarrow{V_R}$$

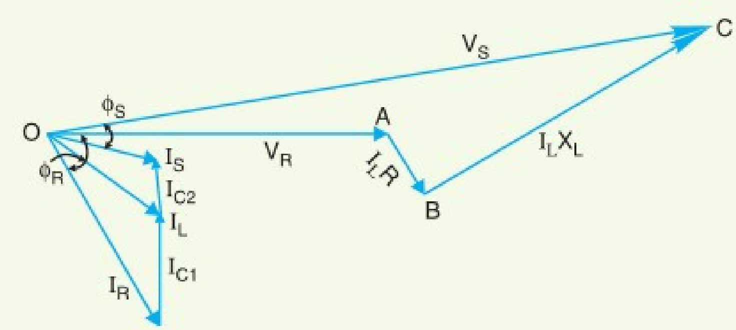


Fig. 10.17

Line current, $\qquad \overrightarrow{I_L} = \overrightarrow{I_R} + \overrightarrow{I_{C1}}$

Sending end voltage, $\qquad \overrightarrow{V_S} = \overrightarrow{V_R} + \overrightarrow{I_L}\,\overrightarrow{Z} = \overrightarrow{V_R} + \overrightarrow{I_L}\,(R + jX_L)$

Charging current at the sending end is

$$\overrightarrow{I_{C2}} = j\,\omega\,(C/2)\,\overrightarrow{V_S} = j\,\pi\,f\,C\,\overrightarrow{V_S}$$

$\therefore$ Sending end current, $\qquad \overrightarrow{I_S} = \overrightarrow{I_L} + \overrightarrow{I_{C2}}$

Note the construction of phasor diagram. $\overrightarrow{V_R}$ is taken as the reference phasor represented by OA. The current $\overrightarrow{I_R}$ lags behind $\overrightarrow{V_R}$ by φ_R. The charging current $\overrightarrow{I_{C1}}$ leads $\overrightarrow{V_R}$ by 90°. The line current $\overrightarrow{I_L}$ is the phasor sum of $\overrightarrow{I_R}$ and $\overrightarrow{I_{C1}}$. The drop $AB = I_L R$ is in phase with $\overrightarrow{I_L}$ whereas drop $BC = I_L X_L$ leads $\overrightarrow{I_L}$ by 90°. Then OC represents the sending end voltage $\overrightarrow{V_S}$. The charging current $\overrightarrow{I_{C2}}$ leads $\overrightarrow{V_S}$ by 90°. Therefore, sending end current $\overrightarrow{I_S}$ is the phasor sum of the $\overrightarrow{I_{C2}}$ and $\overrightarrow{I_L}$. The angle φ_S between sending end voltage V_S and sending end current I_S determines the sending end p.f. $\cos \varphi_S$.

Example 10.13 *A 3-phase, 50Hz, 150 km line has a resistance, inductive reactance and capacitive shunt admittance of 0·1 Ω, 0·5 Ω and 3 × 10^{-6} S per km per phase. If the line delivers 50 MW at 110 kV and 0·8 p.f. lagging, determine the sending end voltage and current. Assume a nominal π circuit for the line.*

Solution. Fig. 10.18 shows the circuit diagram for the line.

Total resistance/phase, $R = 0{\cdot}1 \times 150 = 15\ \Omega$

Total reactance/phase, $X_L = 0{\cdot}5 \times 150 = 75\ \Omega$

Capacitive admittance/phase, $Y = 3 \times 10^{-6} \times 150 = 45 \times 10^{-5}$ S

Receiving end voltage/phase, $V_R = 110 \times 10^3/\sqrt{3} = 63{,}508$ V

Load current, $$I_R = \frac{50 \times 10^6}{\sqrt{3} \times 110 \times 10^3 \times 0{\cdot}8} = 328 \text{ A}$$

$\cos \varphi_R = 0{\cdot}8\ ;\ \sin \varphi_R = 0{\cdot}6$

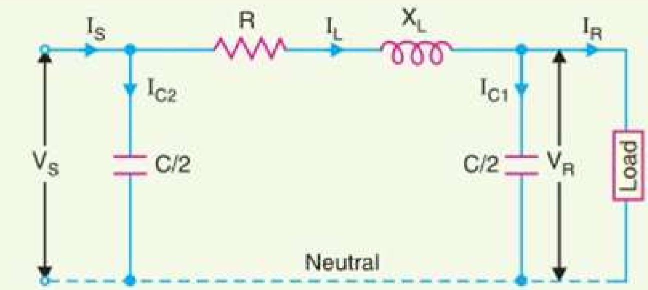


Fig. 10.18

Taking receiving end voltage as the reference phasor, we have,

$$\vec{V_R} = V_R + j\,0 = 63{,}508 \text{ V}$$

Load current, $\vec{I_R} = I_R\,(\cos\varphi_R - j\sin\varphi_R) = 328\,(0{\cdot}8 - j0{\cdot}6) = 262{\cdot}4 - j196{\cdot}8$

Charging current at the load end is

$$\vec{I_{C1}} = \vec{V_R}\,j\,\frac{Y}{2} = 63{,}508 \times j\,\frac{45\times10^{-5}}{2} = j\,14{\cdot}3$$

Line current, $\vec{I_L} = \vec{I_R} + \vec{I_{C1}} = (262{\cdot}4 - j\,196{\cdot}8) + j\,14{\cdot}3 = 262{\cdot}4 - j\,182{\cdot}5$

Sending end voltage,
$$\begin{aligned}\vec{V_S} &= \vec{V_R} + \vec{I_L}\,\vec{Z} = \vec{V_R} + \vec{I_L}\,(R + j\,X_L)\\ &= 63{,}508 + (262{\cdot}4 - j\,182{\cdot}5)\,(15 + j\,75)\\ &= 63{,}508 + 3936 + j\,19{,}680 - j\,2737{\cdot}5 + 13{,}687\\ &= 81{,}131 + j\,16{,}942{\cdot}5 = 82{,}881 \angle\, 11°47' \text{ V}\end{aligned}$$

∴ Line to line sending end voltage = 82,881 × √3 = 1,43,550 V = **143·55 kV**

Charging current at the sending end is

$$\begin{aligned}I_{C2} &= j\vec{V_S}\,Y/2 = (81{,}131 + j\,16{,}942\cdot5)\,j\,\frac{45\times10^{-5}}{2}\\ &= -3{\cdot}81 + j\,18{\cdot}25\end{aligned}$$

Sending end current,
$$\begin{aligned}\vec{I_S} &= \vec{I_L} + \vec{I_{C_2}} = (262{\cdot}4 - j\,182{\cdot}5) + (-3{\cdot}81 + j\,18{\cdot}25)\\ &= 258{\cdot}6 - j\,164{\cdot}25 = 306{\cdot}4 \angle -32{\cdot}4° \text{ A}\end{aligned}$$

∴ Sending end current = **306·4 A**

CHAPTER-5
EHV TRANSMISSION

Explain EHV AC transmission.

EHV AC Transmission line stands for Extra High Voltage Alternating Current transmission line. It can be defined as "A transmission line operating above 220 kV is known as Extra High Voltage AC transmission Line".

Extra-high voltage cables require the power network to go underground due to continuous increase of energy demand, larger transmission capacity, reliability of energy supply and safety.

EHV transmission utilizes increasingly higher voltages for the transmission of. electric power from "high voltage" transmission at 100, 138, 161, and 230 kV to "extra-high voltage" transmission at 345, 400, 500, and 765 kV.

The higher the voltage, the lower the losses. In addition, by being able to move more power across one line, fewer overall lines are needed for transmission.

Explain Reasons for adoption of EHV AC transmission.

1. Reduction of Electrical Losses, Increase in Transmission Efficiency, Improvement of Voltage Regulation and Reduction in Conductor Material Requirement.

2. Line losses are reduced since line losses are inversely proportional to the transmission voltage, Extra high voltage Transmission efficiency increases because of reduction in line losses,

Generating stations (Steam-, hydro- and nuclear-power stations) are located in remote areas (far awayfrom load centers) because of the reasons of economy, feasibility and from the point of view of safety and environmental conditions.

1. The cost of transmission line and terminal equipment also increases with the increase in the transmission voltage but in general these costs are proportional to the transmission voltage rather than the square of the transmission voltage.

Problems involved in EHV transmission.

Limitations:

1. Current Carrying Capacity: The loading of overhead line conductors does not depend on the thermal considerations. However, for overhead transmission lines operating at voltage up to 220 kV,

2. Ferranti Effect: The capacitive load on the line the receiving-end voltage is higher than sending-end voltage.

3. Surge Impedance Loading: Surge impedance loading (SIL) of a transmission line is defined as the load at the receiving end which is transmitted through one end and is also called Natural Load.

4. Mechanical Vibrations and Oscillations: The electrostatic and electromagnetic fields produced by EHV/UHV transmission lines induce currents and voltages.

5. Audible Noise: EHV transmission lines and substations also produce audible noise.

COMPARISION OF DC & AC TRANSMISSION

The electric power can be transmitted either by means of d.c. or a.c. Each system has its own merits and demerits. It is, therefore, desirable to discuss the technical advantages and disadvantages ofthe two systems for transmission of electric power.

1. D.C. transmission. For some years past, the transmission of electric power by d.c. has been receiving the active consideration of engineers due to its numerous advantages.
Advantages. The high voltage d.c. transmission has the following advantages over high voltage a.c. transmission:

(i) It requires only two conductors as compared to three for a.c. transmission.

(ii) There is no inductance, capacitance, phase displacement and surge problems in d.c. transmission.

(iii) Due to the absence of inductance, the voltage drop in a d.c. transmission line is less than the a.c. line for the same load and sending end voltage. For this reason, a d.c. transmission line has better voltage regulation.

(iv) There is no skin effect in a d.c. system. Therefore, entire cross-section of the line conductor is utilized.

(v) For the same working voltage, the potential stress on the insulation is less in case of d.c. system thanthat in a.c. system. Therefore, a d.c. line requires less insulation.

(vi) A d.c. line has less corona loss and reduced interference with communication circuits.

(vii) The high voltage d.c. transmission is free from the dielectric losses, particularly in the case of cables.

(viii) In d.c. transmission, there are no stability problems and synchronizing difficulties.

Disadvantages:

(i) Electric power cannot be generated at high d.c. voltage due to commutation problems.

(ii) The d.c. voltage cannot be stepped up for transmission of power at high voltages.

(iii) The d.c. switches and circuit breakers have their own limitations.

2. A.C. transmission. Now-a-days, electrical energy is almost exclusively generated, transmitted and distributed in the form of a.c.

Advantages.

(i) The power can be generated at high voltages.

(ii) The maintenance of a.c. sub-stations is easy and cheaper.

(iii) The a.c. voltage can be stepped up or stepped down by transformers with ease and efficiency. This permits to transmit power at high voltages and distribute it at safe potentials.

Disadvantages.

(i) An a.c. line requires more copper than a d.c. line.

(ii) The construction of a.c. transmission line is more complicated than a d.c. transmission line.

(iii) Due to skin effect in the a.c. system, the effective resistance of the line is increased.

(iv) An a.c. line has capacitance. Therefore, there is a continuous loss of power due to charging current even when the line is open.

Explain HV DC transmission.

HVDC TRANSMISSION LINE:

From the above comparison, it is clear that high voltage d.c. transmission is superior to high voltage a.c. transmission. Although at present, transmission of electric power is carried by a.c., there is an increasing interest in d.c. transmission. The introduction of mercury arc rectifiers and thyratrons has made it possible to convert a.c. into d.c. and vice-versa easily and efficiently. Such devices can operate upto 30 MW at 400 kV in single units. The present day trend is towards a.c. for generation and distribution and high voltage d.c. for transmission.

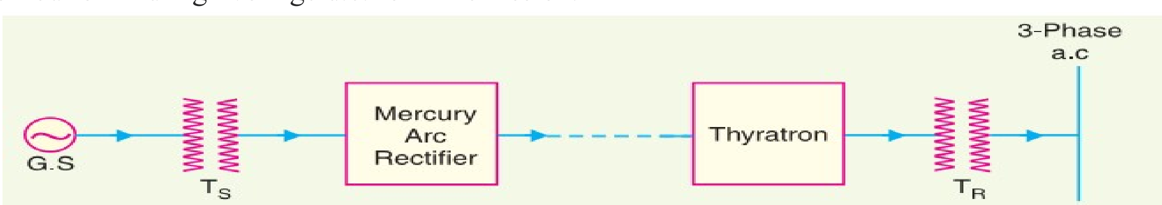


Fig. shows the single line diagram of high voltage d.c. transmission. The electric power is generated as a.c. and is stepped up to high voltage by the sending end transformer T_S. The a.c. power at high voltage is fed to the mercury arc rectifiers which convert a.c. into d.c. The transmission of electric power is carried at high d.c. voltage. At the receiving end, d.c. is converted into a.c. with the help of thyratrons. The a.c. supply is stepped down to low voltage by receiving end transformer T_R for distribution.

State Advantages and Limitations of HVDC transmission system.

ADVANTAGES OF HIGH TRANSMISSION VOLTAGE:

The transmission of electric power is carried at high voltages due to the following reasons :

(i) Reduces volume of conductor material. Consider the transmission of electric power by a three-phase line. Let P = power transmitted in watts

V = line voltage in volts

$$\cos\varphi = \text{power factor of the load}$$
$$l = \text{length of the line in metres}$$
$$R = \text{resistance per conductor in ohms}$$
$$\rho = \text{resistivity of conductor material}$$
$$a = \text{area of X-section of conductor}$$

$$\text{Load current, } I = \frac{P}{\sqrt{3}\,V\cos\varphi}$$

$$\text{Resistance/conductor, } R = \rho l/a$$

$$\text{Total power loss, } W = 3I^2R = 3\left(\frac{P}{\sqrt{3}\,V\cos\varphi}\right)^2 \times \frac{\rho l}{a}$$

$$= \frac{P^2\rho l}{V^2\cos^2\varphi\, a}$$

$$\therefore \quad \text{Area of X-section, } a = \frac{P^2\rho l}{W V^2\cos^2\varphi}$$

Total volume of conductor material required

$$= 3\,a\,l = 3\left(\frac{P^2\rho l}{W V^2\cos^2\varphi}\right) l$$

$$= \frac{3P^2\rho l^2}{W V^2\cos^2\varphi} \qquad \text{.............(i)}$$

It is clear from exp. (i) that for given values of P, l, ρ and W, the volume of conductor material required is inversely proportional to the square of transmission voltage and power factor. In othe words, the greater the transmission voltage, the lesser is the conductor material required.

(ii) Increases transmission efficiency:

$$\text{Input power} = P + \text{Total losses}$$

$$= P + \frac{P^2\rho l}{V^2\cos^2\varphi\, a}$$

Assuming J to be the current density of the conductor, then,

$$a = I/J$$

$$\therefore \quad \text{Input power} = P + \frac{P^2\rho l J}{V^2\cos^2\varphi\, I} = P + \frac{P^2\rho l J}{V^2\cos^2\varphi}\times\frac{1}{I}$$

$$= P + \frac{P^2\rho l J}{V^2\cos^2\varphi}\times\frac{{}^{*}\sqrt{3}\,V\cos\varphi}{P}$$

$$= P + \frac{\sqrt{3}\,P J\rho l}{V\cos\varphi} = P\left[1+\frac{\sqrt{3}\,J\rho l}{V\cos\varphi}\right]$$

$$\text{Transmission efficiency} = \frac{\text{Output power}}{\text{Input power}} = \frac{P}{P\left[1+\frac{\sqrt{3}\,J\rho l}{V\cos\varphi}\right]} = \frac{1}{\left[1+\frac{\sqrt{3}\,J\rho l}{V\cos\varphi}\right]}$$

$$= {}^{**}\left[1-\frac{\sqrt{3}\,J\rho l}{V\cos\varphi}\right] \text{ approx.} \qquad ...(ii)$$

As J, ρ and l are constants, therefore, transmission efficiency increases when the line voltage is increased.

(iii) Decreases percentage line drop:

$$\text{Line drop} = I R = I \times \frac{\rho l}{a}$$

$$= I \times \rho l \times J/I = \rho l J \qquad [\because a = I/J]$$

$$\text{\%age line drop} = \frac{J \rho l}{V} \times 100 \qquad ...(iii)$$

As J, ρ and l are constants, therefore, percentage line drop decreases when the transmission voltage increases.

Limitations of high transmission voltage:-

From the above discussion, it might appear advisable to use the highest possible voltage for transmission of power in a bid to save conductor material. However, it must be realized that high transmission voltage results in

(i) The increased cost of insulating the conductors

(ii) The increased cost of transformers, switchgear and other terminal apparatus.

Therefore, there is a limit to the higher transmission voltage which can be economically employed in a particular case. This limit is reached when the saving in cost of conductor material dueto higher voltage is offset by the increased cost of insulation, transformer, switchgear etc. Hence, the choice of proper transmission voltage is essentially a question of economi

CHAPTER-6

DISTRIBUTION SYSTEMS

Introduction to Distribution System. Explain Connection Schemes of Distribution system (Radial, Ring Main and Inter connected system)

In general, distribution system is that part of power system which distributes power to the consumers for utilization.

That part of power system which distributes electric power for local use is known as distribution system.

In general, the distribution system consists of feeders, distributors and the service mains.

Feeders. A feeder is a conductor which connects the sub-station to the area where power is to be distributed. Generally, no tapings are taken from the feeder so that current in it remains the same throughout. The main consideration in the design of a feeder is the current carrying capacity.

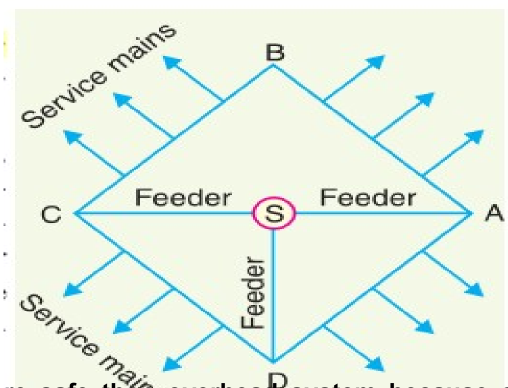
Distributor. A distributor is a conductor from which tapings are taken for supply to the consumers. In Fig. AB, BC, CD and DA are the distributors. The current through a distributor is not constant because tapings are taken at various places along its length.

Service mains. A service main is generally a small cable which connects the distributor to the consumers' terminals.

Classification of distribution system

A distribution system may be classified according to :

- **(i) Nature of current. According to nature of current, distribution system may be classified as (a) d.c. distribution system**
  (b) a.c. distribution system.
- **(ii) Type of construction. According to type of construction, distribution system may be classified as (a) overhead system**
  (b) underground system.
- **(iii) Scheme of connection. According to scheme of connection, the distribution system may be classified as (a) radial system**
  (b) ring main system
  (c) inter-connected system.



Comparison between overhead & underground systems:-

- **(i) Public safety. The underground system is more safe than overhead system because all distribution wiring is placed underground and there are little chances of any hazard.**
- **(ii) Initial cost. The underground system is more expensive due to the high cost of trenching, conduits, cables, manholes and other special equipment. The initial cost of an underground system may be five to ten times than that of an overhead system.**
- **(iii) Flexibility. The overhead system is much more flexible than the underground system. In the latter case, manholes, duct lines etc., are permanently placed once installed and the load expansion can only be met by laying new lines. However, on an overhead system, poles, wires, transformers etc., can be easily shifted to meet the changes in load conditions.**
- **(iv) Faults. The chances of faults in underground system are very rare as the cables are laid underground and are generally provided with better insulation.**
- **(v) Appearance. The general appearance of an underground system is better as all the distribution lines are invisible. This factor is exerting considerable public pressure on electric supply companies to switch over to underground system.**
- **(vi) Fault location and repairs. In general, there are little chances of faults in an underground system. However, if a fault does occur, it is difficult to locate and repair on this system. On an overhead system, the conductors are visible and easily accessible so that fault locations and repairs can be easily made.**
- **(vii) Current carrying capacity and voltage drop. An overhead distribution conductor has a considerably higher current carrying capacity than an underground cable conductor of the same material and cross-section. On the other hand, underground cable conductor has much**

lower inductive reactance than that of an overhead conductor because of closer spacing of conductors.

(viii) Useful life. The useful life of underground system is much longer than that of an overhead system. An overhead system may have a useful life of 25 years, whereas an underground system may have a useful life of more than 50 years.

(ix) Maintenance cost. The maintenance cost of underground system is very low as compared with that of overhead system because of less chance of faults and service interruptions from wind, ice, lightning as well as from traffic hazards.

(x) Interference with communication circuits. An overhead system causes electromagnetic interference with the telephone lines. The power line currents are superimposed on speech currents, resulting in the potential of the communication channel being raised to an undesirable level. However, there is no such interference with the underground system.

<u>Connection scheme of distribution system:-</u>

All distribution of electrical energy is done by constant voltage system. In practice, the following distribution circuits are generally used:

(i) Radial System. In this system, separate feeders radiate from a single substation and feed the distributors at one end only. Fig. (i) Shows a single line diagram of a radial system for d.c. distribution where a feeder OC supplies a distributor AB at point A. Obviously, the distributor is fed at one end only i.e., point A is this case. Fig. (ii) Shows a single line diagram of radial system for a.c. distribution. The radial system is employed only when power is generated at low voltage and the substation is located at the center of the load.

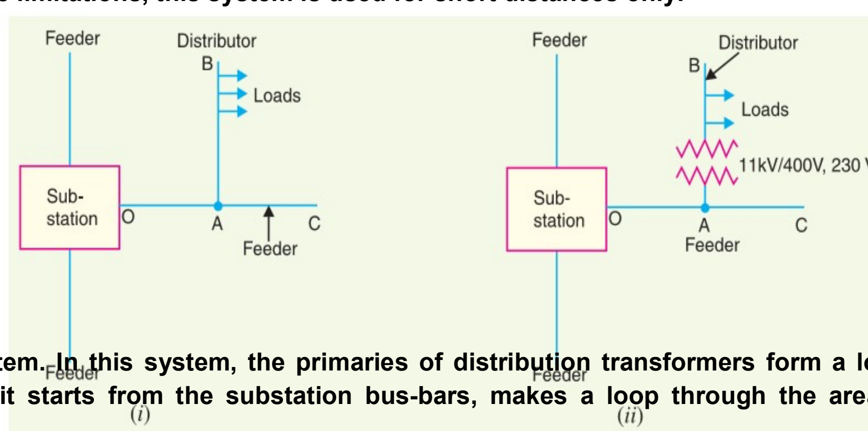
This is the simplest distribution circuit and has the lowest initial cost. However, it suffers from the following drawbacks:

(a) The end of the distributor nearest to the feeding point will be heavily loaded.

(b) The consumers are dependent on a single feeder and single distributor. Therefore, any fault on the feeder or distributor cuts off supply to the consumers who are on the side of the fault away from the substation.

(c) The consumers at the distant end of the distributor would be subjected to serious voltagefluctuations when the load on the distributor changes.

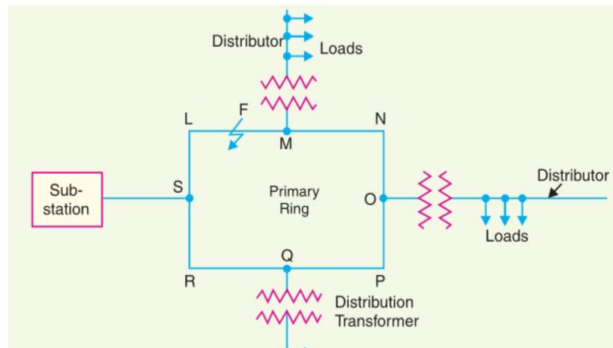
Due to these limitations, this system is used for short distances only.



(ii) Ring main system. In this system, the primaries of distribution transformers form a loop. The loop circuit starts from the substation bus-bars, makes a loop through the area to be served, and returns to the substation. Fig. shows the single line diagram of ring main system for a.c. distribution where substation supplies to the closed feeder LMNOPQRS. The distributors are tapped from different points M, O and Q of the feeder through distribution transformers. The ring main system has the following advantages:

(a) There are less voltage fluctuations at consumer's terminals.

(b) The system is very reliable as each distributor is fed via two feeders. In the event of fault on any section of the feeder, the continuity of supply is maintained. For example, suppose that fault occurs at any point F of section SLM of the feeder. Then section SLM of the feeder can be isolated for repairs and at the same time continuity of supply is maintained to all the consumers via the feeder SRQPONM.

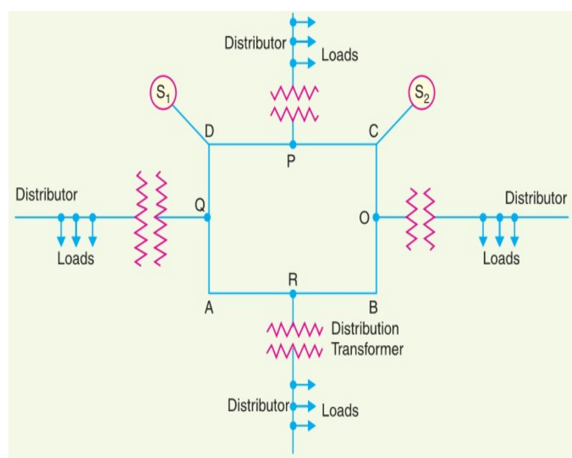


(iii) Interconnected system. When the feeder ring is energized by two or more than two generating stations or substations, it is called inter-connected system. Fig. shows the single line diagram of interconnected system where the closed feeder ring ABCD is supplied by two substations S1 and S2 at points D and C respectively.

Distributors are connected to points O, P, Q and R of the feeder ring through distribution transformers. The interconnected system has the following advantages:

(a) It increases the service reliability.

(b) Any area fed from one generating station during peak load hours can be fed from the other generating station. This reduces reserve power capacity and increases efficiency of the system.



Explain DC distributions (a) Distributor fed at one End (b) Distributor fed atboth the ends (c) Ring distributors.

Now-a-days, electrical energy is generated, transmitted and distributed in the form of a.c. as an economical proposition. However, for certain applications, d.c. supply is absolutely necessary. For example, d.c. supply is required for the operation of variable speed machinery (e.g. d.c. motors), electrochemical work and electric traction. For this purpose, a.c. power is converted into d.c. power at the sub-station by using converting machinery e.g. mercury are rectifiers, rotary converters and motor-generator sets. The d.c. supply from the sub-station is conveyed to the required places for distribution. In this chapter, we shall confine our attention to the various aspects of d.c. distribution.

Types Of DC Distributor

The most general method of classifying d.c. distributors is the way they are fed by the feeders. On this basis, d.c. distributors are classified as:

(i) Distributor fed at one end.
(ii) Distributor fed at both ends.
(iii) Distributor fed at the center.
(iv) Ring distributor.

Distributor fed at one end. In this type of feeding, the distributor is connected to the supply at one end and loads are taken at different points along the length of the distributor. Fig. 13.1 shows the singleline diagram of a d.c. distributor AB fed at the end A (also known as singly fed distributor) and loads I1, I2 and I3 tapped off at points C, D and E respectively.

The following points are worth noting in a singly fed distributor :

(a) The current in the various sections of the distributor away from feeding point goes on decreasing. Thus current in section AC is more than the current in section CD and current in section CD is more than the current in section DE.

(b) The voltage across the loads away from the feeding point goes on decreasing. Thus in Fig. 13.1, the minimum voltage occurs at the load point E.

(c) In case a fault occurs on any section of the distributor, the whole distributor will have to be disconnected from the supply mains. Therefore, continuity of supply is interrupted.

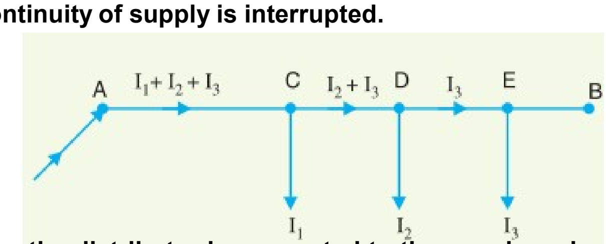


Fig. 13.1

Distributor fed at both ends. In this type of feeding, the distributor is connected to the supply mainsat both ends and loads are tapped off at different points along the length of the distributor. The voltageat the feeding points may or may not be equal. Fig. 13.2 shows a distributor AB fed at the ends A andB and loads of I1, I2 and I3 tapped off at points C, D and E respectively.

Here, the load voltage goes on decreasing as we move away from one feeding point say A, reaches minimum value and then again starts rising and reaches maximum value when we reach the other feeding point B. The minimum voltage occurs at some load point and is never fixed. It is shifted

with the variation of load on different sections of the distributor.

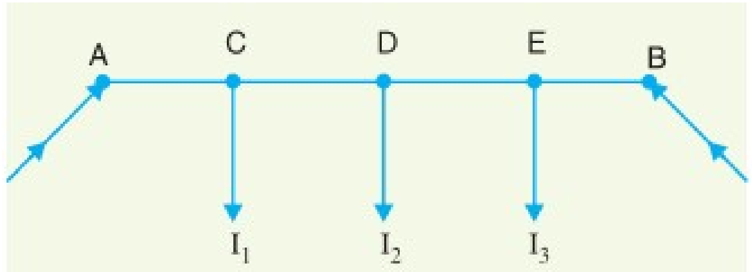


Fig. 13.2

Advantages

(a) If a fault occurs on any feeding point of the distributor, the continuity of supply is maintained from the other feeding point.

(b) In case of fault on any section of the distributor, the continuity of supply is maintained from the other feeding point.

(c) The area of X-section required for a doubly fed distributor is much less than that of a singly fed distributor.

Distributor fed at the Centre. In this type of feeding, the Centre of the distributor is connected to the supply mains as shown in Fig. 13.3. It is equivalent to two singly fed distributors, each distributor having a common feeding point and length equal to half of the total length.

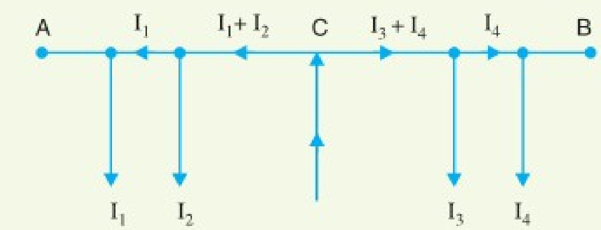


Fig. 13.3

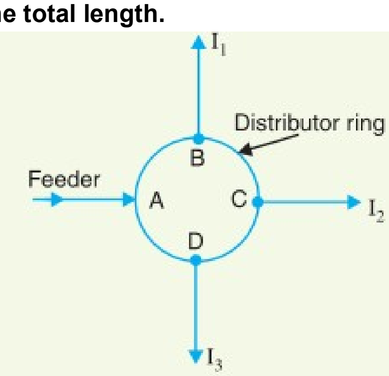


Fig. 13.4

Ring mains. In this type, the distributor is in the form of a closed ring as shown in Fig.13.4. It is equivalent to a straight distributor fed at both ends with equal voltages, the two ends being brought together to form a closed ring. The distributor ring may be fed at one or more than one point.

Explain AC distribution system. Explain Method of solving AC distributionproblem.

The principal disadvantage of d.c. system was that voltage level could not readily be changed, except by the use of rotating machinery, which in most cases was too expensive.

Now-a-days, electrical energy is generated, transmitted and distributed in the form of alternating current as an economical proposition. The electrical energy produced at the power station is transmitted at very high voltages by 3-phase, 3wire system to step-down sub-stations for distribution. The distribution system consists of two parts viz. primary distribution and secondary distribution.

The primary distribution circuit is 3phase, 3-wire and operates at voltages (3·3 or 6·6 or 11kV) somewhat higher than general utilization levels. It delivers power to the secondary distribution circuit through distribution transformers situated near consumers' localities. Each distribution transformer steps down the voltage to 400 V and power is distributed to ultimate consumers' by 400/230 V, 3-phase, 4-wire system.

AC Distribution Calculations

A.C. distribution calculations differ from those of d.c. distribution in the following respects :

(i) In case of d.c. system, the voltage drop is due to resistance alone. However, in a.c. system, the voltage drops are due to the combined effects of resistance, inductance and capacitance.

(ii) In a d.c. system, additions and subtractions of currents or voltages are done arithmetically but in case of a.c. system, these operations are done vectorially.

(iii) In an a.c. system, power factor has to be taken into account.Loads tapped off form the distributor are generally at different power factors. There are two ways of referring power factor

(a) It may be referred to supply or receiving end voltage which is regarded as the reference vector.

(b) It may be referred to the voltage at the load point itself.

There are several ways of solving a.c. distribution problems. However, symbolic notation method has been found to be most convenient for this purpose. In this method, voltages, currents and impedances are expressed in complex notation and the calculations are made exactly as in d.c. distribution.

Methods Of Solving AC Distribution Problems :-

In a.c. distribution calculations, power factors of various load currents have to be considered since currents in different sections of the distributor will be the vector sum of load currents and not the arithmetic sum. The power factors of load currents may be given (i) w.r.t. receiving or sending end voltage or (ii) w.r.t. to load voltage itself. Each case shall be discussed separately.

(i) Power factors referred to receiving end voltage. Consider an a.c. distributor AB with concentrated loads of I_1 and I_2 tapped off at points C and B as shown in Fig. 14.1. Taking the receivingend voltage V_B as the reference vector, let lagging power factors at C and B be cos φ_1 and cos φ_2 w.r.t.V_B. Let R_1, X_1 and R_2, X_2 be the resistance and reactance of sections AC and CB of the distributor.

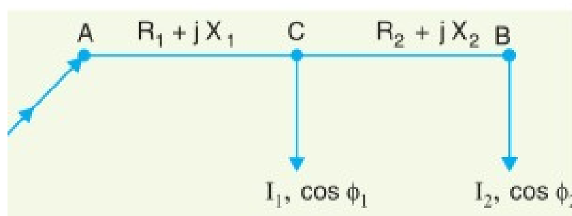


Fig. 14.1

Impedance of section AC, $\overrightarrow{Z_{AC}} = R_1 + j X_1$

Impedance of section CB, $\overrightarrow{Z_{CB}} = R_2 + j X_2$

Load current at point C, $\overrightarrow{I_1} = I_1 (\cos \varphi_1 - j \sin \varphi_1)$

Load current at point B, $\overrightarrow{I_2} = I_2 (\cos \varphi_2 - j \sin \varphi_2)$

Current in section CB, $\overrightarrow{I_{CB}} = \overrightarrow{I_2} = I_2 (\cos \varphi_2 - j \sin \varphi_2)$

Current in section AC, $\overrightarrow{I_{AC}} = \overrightarrow{I_1} + \overrightarrow{I_2}$

$$= I_1 (\cos \varphi_1 - j \sin \varphi_1) + I_2 (\cos \varphi_2 - j \sin \varphi_2)$$

Voltage drop in section CB, $\overrightarrow{V_{CB}} = \overrightarrow{I_{CB}}\,\overrightarrow{Z_{CB}} = I_2 (\cos \varphi_2 - j \sin \varphi_2)(R_2 + j X_2)$

Voltage drop in section AC, $\overrightarrow{V_{AC}} = \overrightarrow{I_{AC}}\,\overrightarrow{Z_{AC}} = \left(\overrightarrow{I_1} + \overrightarrow{I_2}\right) Z_{AC}$

$$= [I_1(\cos \varphi_1 - j \sin \varphi_1) + I_2 (\cos \varphi_2 - j \sin \varphi_2)]\,[R_1 + jX_1]$$

Sending end voltage, $\overrightarrow{V_A} = \overrightarrow{V_B} + \overrightarrow{V_{CB}} + \overrightarrow{V_{AC}}$

Sending end current, $\overrightarrow{I_A} = \overrightarrow{I_1} + \overrightarrow{I_2}$

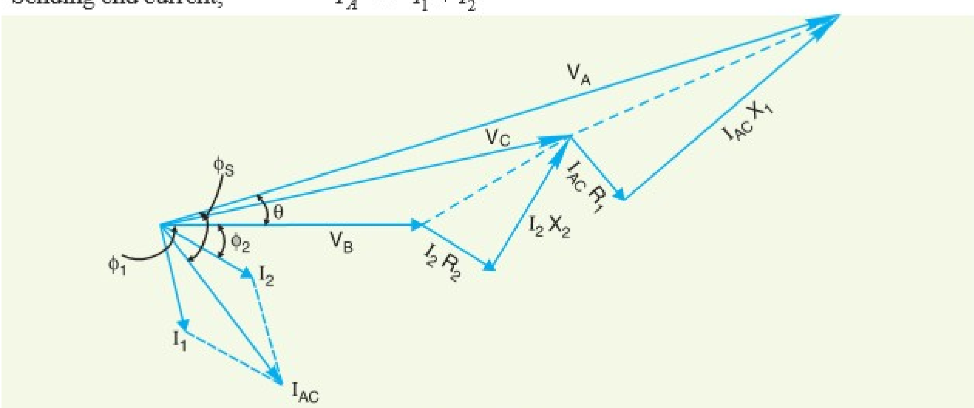


Fig. 14.2

The vector diagram of the a.c. distributor under these conditions is shown in Fig. 14.2. Here, the receiving end voltage VB is taken as the reference vector. As power factors of loads are given w.r.t. VB, therefore, I1 and I2 lag behind VB by φ1 and φ2 respectively.

(ii) Power factors referred to respective load voltages. Suppose the power factors of loads in the previous Fig. 14.1 are referred to their respective load voltages. Then φ1 is the phase angle between VC and I1 and φ2 is the phase angle between VB and I2. The vector diagram under these conditions is shown in Fig. 14.3.

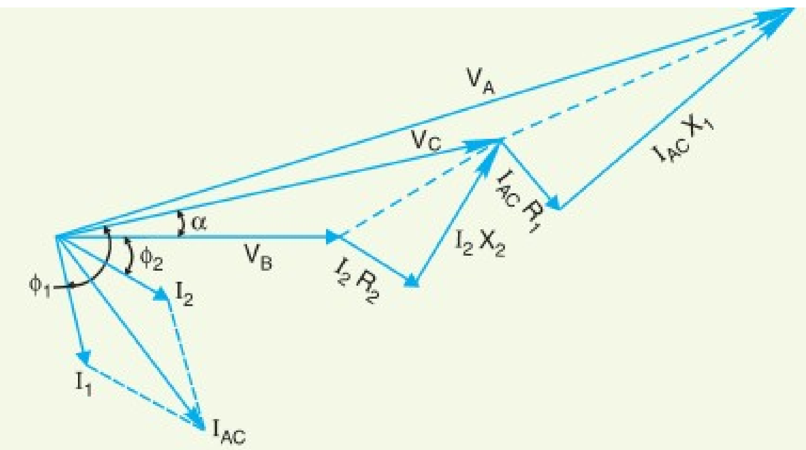


Fig. 14.3

$$\text{Voltage drop in section } CB = \overrightarrow{I_2}\,\overrightarrow{Z_{CB}} = I_2(\cos\varphi_2 - j\sin\varphi_2)(R_2 + j X_2)$$

$$\text{Voltage at point } C = \overrightarrow{V_B} + \text{Drop in section } CB = V_C \angle \alpha \text{ (say)}$$

Now $\overrightarrow{I_1} = I_1 \angle -\varphi_1$ *w.r.t.* voltage V_C

$\therefore$ $\overrightarrow{I_1} = I_1 \angle -(\varphi_1 - \alpha)$ *w.r.t.* voltage V_B

i.e. $\overrightarrow{I_1} = I_1 [\cos(\varphi_1 - \alpha) - j\sin(\varphi_1 - \alpha)]$

Now $\overrightarrow{I_{AC}} = \overrightarrow{I_1} + \overrightarrow{I_2}$

$$= I_1[\cos(\varphi_1 - \alpha) - j\sin(\varphi_1 - \alpha)] + I_2(\cos\varphi_2 - j\sin\varphi_2)$$

$$\text{Voltage drop in section } AC = \overrightarrow{I_{AC}}\,\overrightarrow{Z_{AC}}$$

$\therefore$ Voltage at point $A = V_B$ + Drop in CB + Drop in AC

Example 14.1. *A single phase a.c. distributor AB 300 metres long is fed from end A and is loaded as under :*

(i) 100 A at 0·707 p.f. lagging 200 m from point A

(ii) 200 A at 0·8 p.f. lagging 300 m from point A

The load resistance and reactance of the distributor is 0·2 Ω and 0·1 Ω per kilometre. Calculate the total voltage drop in the distributor. The load power factors refer to the voltage at the far end.

Solution. Fig. 14.4 shows the single line diagram of the distributor.

Impedance of distributor/km $= (0{\cdot}2 + j\,0{\cdot}1)\ \Omega$

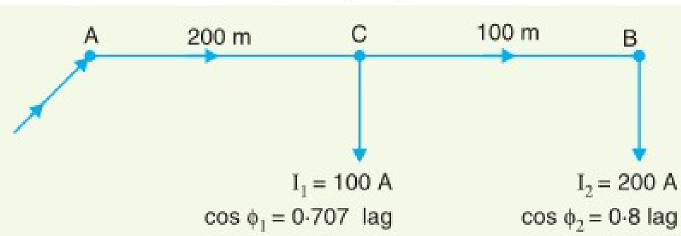


Fig. 14.4

Impedance of section AC, $\overrightarrow{Z_{AC}} = (0{\cdot}2 + j\,0{\cdot}1) \times 200/1000 = (0{\cdot}04 + j\,0{\cdot}02)\ \Omega$

Impedance of section CB, $\overrightarrow{Z_{CB}} = (0{\cdot}2 + j\,0{\cdot}1) \times 100/1000 = (0{\cdot}02 + j\,0{\cdot}01)\ \Omega$

Taking voltage at the far end B as the reference vector, we have,

Load current at point B, $\vec{I_2} = I_2(\cos\phi_2 - j\sin\phi_2) = 200\,(0{\cdot}8 - j\,0{\cdot}6)$
$= (160 - j\,120)$ A

Load current at point C, $\vec{I_1} = I_1(\cos\phi_1 - j\sin\phi_1) = 100\,(0{\cdot}707 - j\,0{\cdot}707)$
$= (70{\cdot}7 - j\,70{\cdot}7)$ A

Current in section CB, $\overrightarrow{I_{CB}} = \vec{I_2} = (160 - j\,120)$ A

Current in section AC, $\overrightarrow{I_{AC}} = \vec{I_1} + \vec{I_2} = (70{\cdot}7 - j\,70{\cdot}7) + (160 - j\,120)$
$= (230{\cdot}7 - j\,190{\cdot}7)$ A

Voltage drop in section CB, $\overrightarrow{V_{CB}} = \overrightarrow{I_{CB}}\,\overrightarrow{Z_{CB}} = (160 - j\,120)\,(0{\cdot}02 + j\,0{\cdot}01)$
$= (4{\cdot}4 - j\,0{\cdot}8)$ volts

Voltage drop in section AC, $\overrightarrow{V_{AC}} = \overrightarrow{I_{AC}}\,\overrightarrow{Z_{AC}} = (230{\cdot}7 - j\,190{\cdot}7)\,(0{\cdot}04 + j\,0{\cdot}02)$
$= (13{\cdot}04 - j\,3{\cdot}01)$ volts

Voltage drop in the distributor $= \overrightarrow{V_{AC}} + \overrightarrow{V_{CB}} = (13{\cdot}04 - j\,3{\cdot}01) + (4{\cdot}4 - j\,0{\cdot}8)$
$= (17{\cdot}44 - j\,3{\cdot}81)$ volts

Magnitude of drop $= \sqrt{(17\cdot44)^2 + (3\cdot81)^2}$ = **17·85 V**

Example 14.2. *A single phase distributor 2 kilometres long supplies a load of 120 A at 0·8 p.f. lagging at its far end and a load of 80 A at 0·9 p.f. lagging at its mid-point. Both power factors are referred to the voltage at the far end. The resistance and reactance per km (go and return) are 0·05 Ω and 0·1 Ω respectively. If the voltage at the far end is maintained at 230 V, calculate :*

(i) voltage at the sending end

(ii) phase angle between voltages at the two ends.

Solution. Fig. 14.5 shows the distributor AB with C as the mid-point

Impedance of distributor/km $= (0{\cdot}05 + j\,0{\cdot}1)\ \Omega$

Impedance of section AC, $\vec{Z}_{AC} = (0{\cdot}05 + j\,0{\cdot}1) \times 1000/1000 = (0{\cdot}05 + j\,0{\cdot}1)\ \Omega$

Impedance of section CB, $\vec{Z}_{CB} = (0{\cdot}05 + j\,0{\cdot}1) \times 1000/1000 = (0{\cdot}05 + j\,0{\cdot}1)\ \Omega$

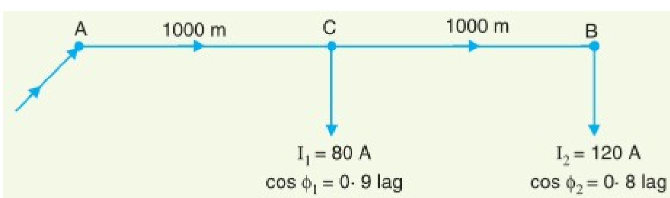


Fig. 14.5

Let the voltage V_B at point B be taken as the reference vector.

Then, $\overrightarrow{V_B} = 230 + j\,0$

(*i*) Load current at point B, $\overrightarrow{I_2} = 120\,(0{\cdot}8 - j\,0{\cdot}6) = 96 - j\,72$

Load current at point C, $\overrightarrow{I_1} = 80\,(0{\cdot}9 - j\,0{\cdot}436) = 72 - j\,34{\cdot}88$

Current in section CB, $\overrightarrow{I_{CB}} = \overrightarrow{I_2} = 96 - j\,72$

Current in section AC,
$$\overrightarrow{I_{AC}} = \overrightarrow{I_1} + \overrightarrow{I_2} = (72 - j\,34{\cdot}88) + (96 - j\,72)$$
$$= 168 - j\,106{\cdot}88$$

Drop in section CB,
$$\overrightarrow{V_{CB}} = \overrightarrow{I_{CB}}\,\overrightarrow{Z_{CB}} = (96 - j\,72)\,(0{\cdot}05 + j\,0{\cdot}1)$$
$$= 12 + j\,6$$

Drop in section AC,
$$\overrightarrow{V_{AC}} = \overrightarrow{I_{AC}}\,\overrightarrow{Z_{AC}} = (168 - j\,106{\cdot}88)\,(0{\cdot}05 + j\,0{\cdot}1)$$
$$= 19{\cdot}08 + j\,11{\cdot}45$$

∴ Sending end voltage,
$$\overrightarrow{V_A} = \overrightarrow{V_B} + \overrightarrow{V_{CB}} + \overrightarrow{V_{AC}}$$
$$= (230 + j\,0) + (12 + j\,6) + (19.08 + j\,11.45)$$
$$= 261.08 + j\,17.45$$

Its magnitude is $= \sqrt{(261{\cdot}08)^2 + (17{\cdot}45)^2} = \mathbf{261{\cdot}67\ V}$

(*ii*) The phase difference θ between V_A and V_B is given by :

$$\tan\theta = \frac{17{\cdot}45}{261{\cdot}08} = 0{\cdot}0668$$

∴ $\theta = \tan^{-1} 0{\cdot}0668 = \mathbf{3{\cdot}82^{\circ}}$

Explain three phase four wire star connected system arrangement.

The 3-phase loads that have the same impedance and power factor in each phase are called balanced loads. The problems on balanced loads can be solved by considering one phase only; the conditions in the other two phases being similar. However, we may come across a situation when loads are unbalanced i.e. each load phase has different impedance and/or power factor. In that case, current and power in each phase will be different. In practice, we may come across the following unbalanced loads :

(i) Four-wire star-connected unbalanced load

(ii) Unbalanced Δ-connected load

(iii) Unbalanced 3-wire, Y-connected load

The 3-phase, 4-wire system is widely used for distribution of electric power in commercial and industrial buildings. The single phase load is connected between any line and neutral wire while a 3-phase load is connected across the three lines. The 3-phase, 4-wire system invariably carries unbalanced loads.

Three phase four wire star connected Unbalanced Loads

We can obtain this type of load in two ways. First, we may connect a 3-phase, 4-wire unbalanced load to a 3-phase, 4-wire supply as shown in Fig. 14.10. Note that star point N of the

supply is connected to the load star point N′. Secondly, we may connect single phase loads between any line and the neutral wire as shown in Fig.14.11. This will also result in a 3-phase, 4-wire **unbalanced load because it is rarely possible that single phase loads on all the three phases have the same magnitude and power factor. Since the load is unbalanced, the line currents will be different in magnitude and displaced from one another by unequal angles. The current in the neutral wire will be the phasor sum of the three line currents i.e.

Current in neutral wire, $I_N = I_R + I_Y + I_B$ .... phasor sum

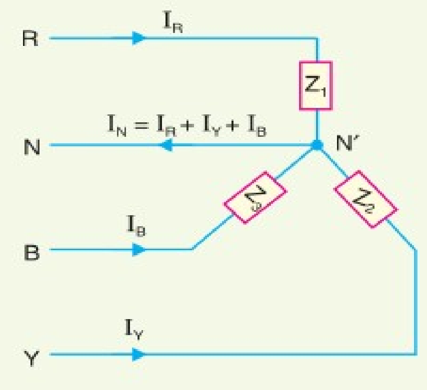


Fig. 14.10

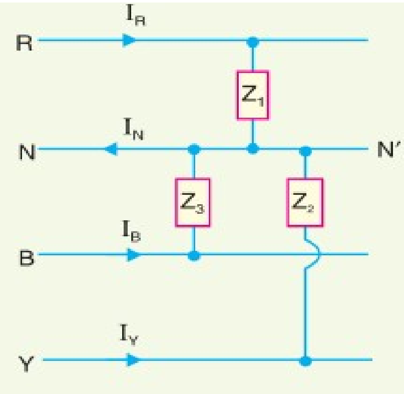


Fig. 14.11

The following points may be noted carefully :

(i) Since the neutral wire has negligible resistance, supply neutral N and load neutral N′ will be at the same potential. It means that voltage across each impedance is equal to the phase voltage of the supply.However, current in each phase (or line) will be different due to unequal impedances.

(ii) The amount of current flowing in the neutral wire will depend upon the magnitudes of line currents and their phasor relations. In most circuits encountered in practice, the neutral current is equal to or smaller than one of the line currents. The exceptions are those circuits having severe unbalance.

Example 14.7. *Non-reactive loads of 10 kW, 8 kW and 5 kW are connected between the neutral and the red, yellow and blue phases respectively of a 3-phase, 4-wire system. The line voltage is 400V. Calculate (i) the current in each line and (ii) the current in the neutral wire.*

Solution. This is a case of unbalanced load so that the line currents (and hence the phase currents) in the three lines will be different. The current in the *neutral wire will be equal to the phasor sum of three line currents as shown in Fig. 14.12.

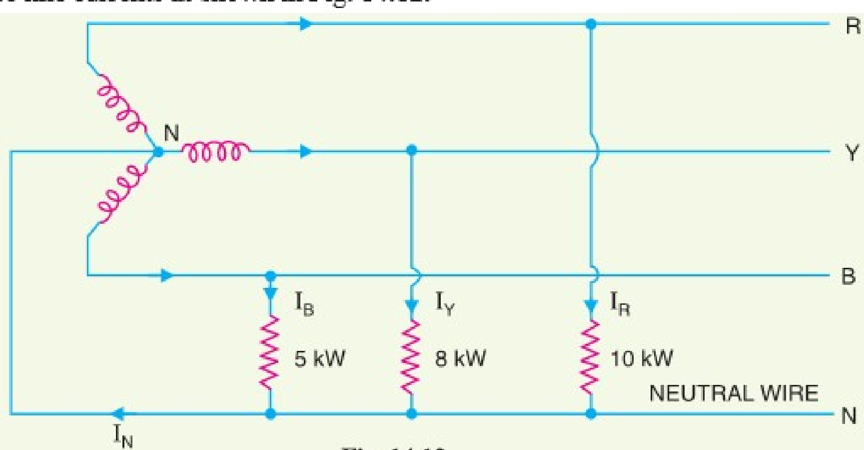


Fig. 14.12

(i) Phase voltage $= 400/\sqrt{3} = 231$ V

$$I_R = 10 \times 10^3/231 = \mathbf{43{\cdot}3\ A}$$

$$I_Y = 8 \times 10^3/231 = \mathbf{34{\cdot}6\ A}$$

$$I_B = 5 \times 10^3/231 = \mathbf{21{\cdot}65\ A}$$

(ii) The three lines currents are represented by the respective phasors in Fig. 14.13. Note that the three line currents are of different magnitude but displaced 120° from one another. The current in the neutral wire will be the phasor sum of the three line currents.

Resolving the three currents along x-axis and y-axis, we have,

Resultant horizontal component $= I_Y \cos 30° - I_B \cos 30°$

$= 34{\cdot}6 \times 0{\cdot}866 - 21{\cdot}65 \times 0{\cdot}866 = 11{\cdot}22$ A

Resultant vertical component $= I_R - I_Y \cos 60° - I_B \cos 60°$

$= 43{\cdot}3 - 34{\cdot}6 \times 0{\cdot}5 - 21{\cdot}65 \times 0{\cdot}5 = 15{\cdot}2$ A

As shown in Fig. 14.14, current in neutral wire is

$$I_N = \sqrt{(11{\cdot}22)^2 + (15{\cdot}2)^2} = \mathbf{18{\cdot}9\ A}$$

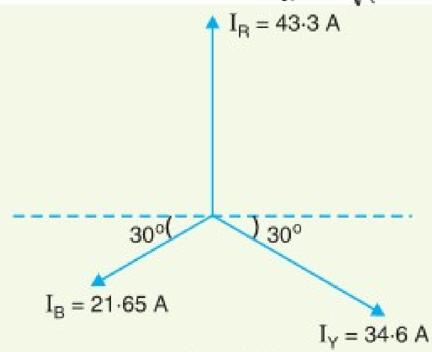


Fig. 14.13

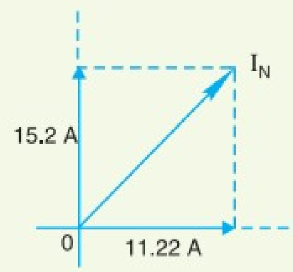


Fig. 14.14

CHAPTER-7
UNDERGROUND CABLES

a. Explain cable insulation and classification of cables.

Electric power can be transmitted or distributed either by overhead system or by underground cables. The underground cables have several advantages such as less liable to damage through storms or lightning, low maintenance cost, less chance of faults, smaller voltage drop and better general appearance. However, their major drawback is that they have greater installation cost and introduce insulation problems at high voltages compared with the equivalent overhead system.

"An underground cable essentially consists of one or more conductors covered with suitable insulation and surrounded by a protecting cover."

The use of underground cables for many years has been for distribution of electric power in congested urban areas at comparatively low or moderate voltages.

Although several types of cables are available, the type of cable to be used will depend upon the working voltage and service requirements. In general, a cable must fulfill the following necessary requirements:

(i) The conductor used in cables should be tinned stranded copper or aluminum of high conductivity. Stranding is done so that conductor may become flexible and carry more current.

(ii)The conductor size should be such that the cable carries the desired load current without overheatingand causes voltage drop within permissible limits.

(iii)The cable must have proper thickness of insulation in order to give high degree of safety and

reliability at the voltage for which it is designed.

(iv)The cable must be provided with suitable mechanical protection so that it may withstand the rough use in laying it.

(v)The materials used in the manufacture of cables should be such that there is complete chemical and physical stability throughout.

Construction of cable

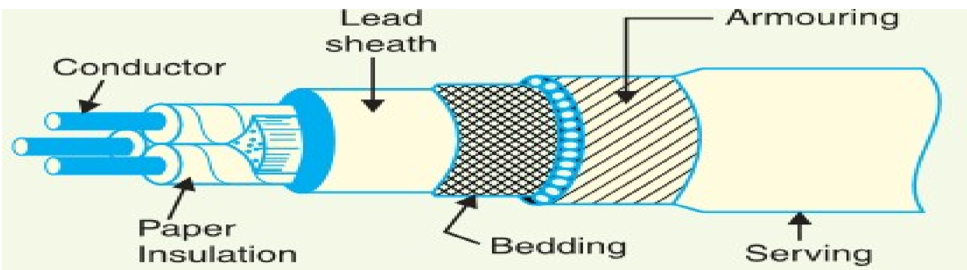
Fig. shows the general construction of a 3-conductor cable. The various parts are:

(i) Cores or Conductors. A cable may have one or more than one core (conductor) depending upon the type of service for which it is intended. For instance, the 3-conductor cable shown in Fig. is used for 3-phase service. The conductors are made of tinned copper or aluminum and are usually stranded in order to provide flexibility to the cable.

(ii) Insulation. Each core or conductor is provided with a suitable thickness of insulation, the thickness of layer depending upon the voltage to be withstood by the cable. The commonly used materials for insulation are impregnated paper, varnished cambric or rubber mineral compound.

(iii) Metallic sheath. In order to protect the cable from moisture, gases or other damaging liquids (acids or alkalis) in the soil and atmosphere, a metallic sheath of lead or aluminum is provided over the insulation as shown in Fig.

(iv) Bedding. Over the metallic sheath is applied a layer of bedding which consists of a fibrous material like jute or hessian tape. The purpose of bedding is to protect the metallic sheath against corrosion andfrom mechanical injury due to armoring.



(v) Armouring. Over the bedding, armouring is provided which consists of one or two layers of galvanised steel wire or steel tape. Its purpose is to protect the cable from mechanical injury while laying it and during the course of handling. Armouring may not be done in the case of some cables.

(vi) Serving. In order to protect armouring from atmospheric conditions, a layer of fibrous material (like jute) similar to bedding is provided over the armouring. This is known as serving. **INSULATING MATERIAL FOR CABLES:-**

In general, the insulating materials used in cables should have the following properties :

(i) High insulation resistance to avoid leakage current.

(ii) High dielectric strength to avoid electrical breakdown of the cable.

(iii) High mechanical strength to withstand the mechanical handling of cables.

(iv) Non-hygroscopic i.e., it should not absorb moisture from air or soil. The moisture tends to decrease the insulation resistance and hastens the breakdown of the cable. In case the insulating material is hygroscopic, it must be enclosed in a waterproof covering like lead sheath.

(v) Non-inflammable.

(vi) Low cost so as to make the underground system a viable proposition.

(Vii) Unaffected by acids and alkalies to avoid any chemical action.

b. State Types of L. T. & H.T. cables with constructional features.

CLASSIFICATION OF CABLE:-

Cables for underground service may be classified in two ways according to (i) the type of insulating material used in their manufacture (ii) the voltage for which they are manufactured. However, the latter method of classification is generally preferred, according to which cables can be divided into the following groups :

(i) Low-tension (L.T.) cables — upto 1000 V
(ii) High-tension (H.T.) cables — upto 11,000 V
(iii) Super-tension (S.T.) cables — from 22 kV to 33 kV
(iv) Extra high-tension (E.H.T.) cables — from 33 kV to 66 kV
(v) Extra super voltage cables — beyond 132 kV

A cable may have one or more than one core depending upon the type of service for which it is intended.It may be (i) single-core (ii) two-core (iii) three-core (iv) four-core etc. For a 3-phase service, either 3-single-core cables or three-core cable can be used depending upon the operating voltage and load demand.

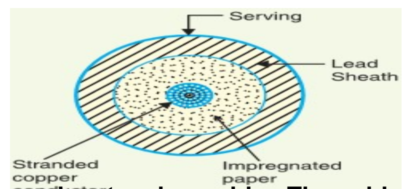


Fig. 11.2 shows the constructional details of a single-core low tension cable. The cable has ordinary construction because the stresses developed in the cable for low voltages (upto 6600 V) are generally small. It consists of one circular core of tinned stranded copper (or aluminum) insulated by layers of impregnated paper. The insulation is surrounded by a lead sheath which prevents the entry of moisture into the inner parts. In order to protect the lead sheath from corrosion, an overall serving of compounded fibrous material (jute etc.) is provided. Single-core cables are not usually armoured in order to avoid excessive sheath losses. The principal advantages of single-core cables are simple construction and availability of larger copper section.

Cable For 3-Phase Service

In practice, underground cables are generally required to deliver 3-phase power. For the purpose, either three-core cable or *three single core cables may be used. For voltages upto 66 kV, 3-core cable (i.e., multi-core construction) is preferred due to economic reasons. However, for voltages beyond 66 kV, 3-core-cables become too large and unwieldy and, therefore, single-core cables are used. The following types of cables are generally used for 3-phase service :

1. Belted cables — upto 11 kV
2. Screened cables — from 22 kV to 66 kV
3. Pressure cables — beyond 66 kV.

Belted cables. These cables are used for voltages upto 11kV but in extraordinary cases, their use may be extended upto 22kV. Fig. 11.3 shows the constructional details of a 3-core belted cable. The cores are insulated from each other by layers of impregnated paper. Another layer of impregnated paper tape, called paper belt is wound round the grouped insulated cores. The gap between the insulated cores is filled with fibrous insulating material (jute etc.) so as to give circular cross-section to the cable. The cores are generally stranded and may be of noncircular shape to make better use of available space. The belt is covered with lead sheath to protect the cable against ingress of moisture and mechanical injury. The lead sheath is covered with one or more layers of armouring with an outer serving.

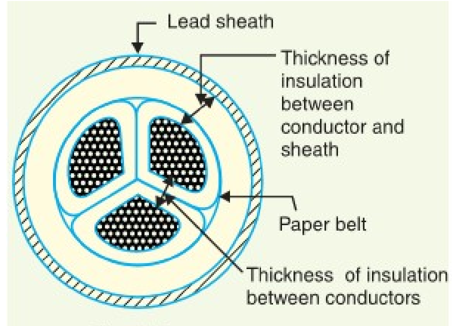


Fig. 11.3

Screened cables. These cables are meant for use upto 33 kV, but in particular cases their use may be extended to operating voltages upto 66 kV. Two principal types of screened cables are Htype cables and S.L. type cables.

(i) H-type cables. This type of cable was first designed by H. Hochstadter and hence the name. Fig. 11.4 shows the constructional details of a typical 3-core, H-type cable. Each core is insulated by layers of impregnated paper. The insulation on each core is covered with a metallic screen which usually consists of a perforated aluminum foil. The cores are laid in such a way that metallic screens make contact with one another. An additional conducting belt (copper woven fabric tape) is wrapped round the three cores. The cable has no insulating belt but lead sheath, bedding, armouring and serving follow as usual. It is easy to see that each core screen is in electrical contact with the conducting belt and the lead sheath. As all the four screens (3 core screens and one conducting belt) and the lead sheath are at †earth potential, therefore, the electrical stresses are purely radial and consequently dielectric losses are reduced. Two principal advantages are claimed for H-type cables. Firstly, the perforations in the metallic screens assist in the complete impregnation of the cable with the compound and thus the possibility of air pockets or voids (vacuous spaces) in the dielectric is eliminated. The voids if present tend to reduce the breakdown strength of the cable and may cause considerable damage to the paper insulation. Secondly, the metallic screens increase the heat dissipating power of the cable.

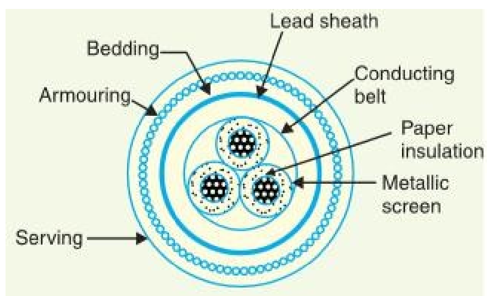


Fig. 11.4

(ii) S.L. type cables. Fig. 11.5 shows the constructional details of a 3-core *S.L. (separate lead) type cable. It is basically H-type cable but the screen round each core insulation is covered by its own lead

sheath. There is no overall lead sheath but only armouring and serving are provided. The S.L. type cables have two main advantages over H-type cables. Firstly, the separate sheaths minimise the possibility of core-to-core breakdown. Secondly, bending of cables becomes easy due to the elimination of overall lead sheath. However, the disadvantage is that the three lead sheaths of S.L. cable are much thinner than the single sheath of H-cable and, therefore, call for greater care in manufacture.

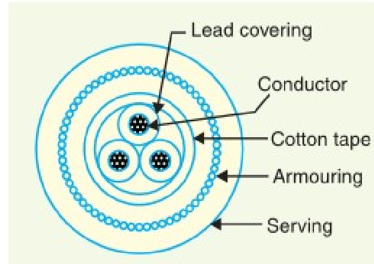


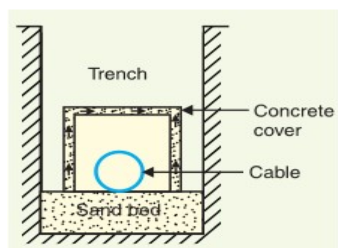
Fig. 11.5

Pressure cables. For voltages beyond 66 kV, solid type cables are unreliable because there is a danger of breakdown of insulation due to the presence of voids. When the operating voltages are greater than 66 kV, pressure cables are used. In such cables, voids are eliminated by increasing the pressure of compound and for this reason they are called pressure cables. Two types of pressure cables via oil-filled cables and gas pressure cables are commonly used.

c. State and Explain Methods of cable lying.

The reliability of underground cable network depends to a considerable extent upon the proper laying and attachment of fittings i.e., cable end boxes, joints, branch connectors etc. There are three main methods of laying underground cables viz., direct laying, draw-in system and the solid system.

Direct laying. This method of laying underground cables is simple and cheap and is much favoured in modern practice. In this method, a trench of about 1·5 metres deep and 45 cm wide is dug. The trench is covered with a layer of fine sand (of about 10 cm thickness) and the cable islaid over this sand bed. The sand prevents the entry of moisture from the ground and thus protects the cable from decay. After the cable has been laid in the trench, it is covered with another layer of sand of about 10 cm thickness. The trench is then covered with bricks and other materials in order to protect the cable from mechanical injury. When more than one cable is to be laid in the same trench, a horizontal or vertical interaxial spacing of atleast 30 cm is provided in order to reduce the effect of mutual heating and also to ensure that a fault occurring on one cable does not damage the adjacent cable. Cables to be laid in this way must have serving of bituminised paper and hessian tape so as to provide protection against corrosion and electorlysis.



Advantages

(i) It is a simple and less costly method.

(ii) It gives the best conditions for dissipating the heat generated in the cables.

(iii) It is a clean and safe method as the cable is invisible and free from external disturbances.

Disadvantages

(i) The extension of load is possible only by a completely new excavation which may cost as much as the original work.

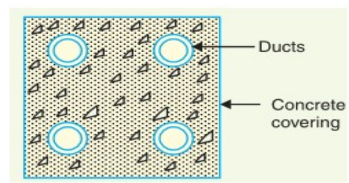
(ii) The alterations in the cable network cannot be made easily.

(iii) The maintenance cost is very high.

(iv) Localization of fault is difficult.

(v) It cannot be used in congested areas where excavation is expensive and inconvenient.

Draw-in system. In this method, conduit or duct of glazed stone or cast iron or concrete are laid in the ground with manholes at suitable positions along the cable route. The cables are then pulled into position from manholes. Fig. shows section through four-way underground duct line. Three of the ducts carry transmission cables and the fourth duct carries relay protection connection, pilot wires. Care must be taken that where the duct line changes direction; depths, dips and offsets be made with a very long radius or it will be difficult to pull a large cable between the manholes. The distance between the manholes should not be too long so as to simplify the pulling in of the cables. The cables to be laid in this way need not be armoured but must be provided with serving of hessian and jute in order to protect them when being pulled into the ducts.



Advantages

(i) Repairs, alterations or additions to the cable network can be made without opening the ground.

(ii) As the cables are not armoured, therefore, joints become simpler and maintenance cost is reducedconsiderably.

(iii) There are very less chances of fault occurrence due to strong mechanical protection provided by the system.

Disadvantages

(i) The initial cost is very high.

(ii) The current carrying capacity of the cables is reduced due to the close grouping of cables and unfavourable conditions for dissipation of heat.

This method of cable laying is suitable for congested areas where excavation is expensive and inconvenient, for once the conduits have been laid, repairs or alterations can be made without opening the ground. This method is generally used for short length cable routes such as in workshops, road crossings where frequent digging is costlier or impossible.

Solid system. In this method of laying, the cable is laid in open pipes or troughs dug out in earth along the cable route. The troughing is of cast iron, stoneware, asphalt or treated wood. After the cable is laid in position, the troughing is filled with a bituminous or asphaltic compound and covered over. Cables laid in this manner are usually plain lead covered because troughing affords good mechanical protection.

Disadvantages

(i) It is more expensive than direct laid system.

(ii) It requires skilled labour and favorable weather conditions.

(iii) Due to poor heat dissipation facilities, the current carrying capacity of the cable is reduced.

In view of these disadvantages, this method of laying underground cables is rarely used now-days.

d. State methods of Localisation of cable faults – Murray and Varley looptest for short circuit fault/Earth fault.

LOOP TEST FOR LOCATION OF FAULTS IN UNDERGROUND CABLES:-

There are several methods for locating the faults in underground cables. However, two popula methods known as loop tests are:

(i) Murray loop test

(ii) Varley loop test

These simple tests can be used to locate the earth fault or short-circuit fault in underground cables provided that a sound cable runs along the faulty cable. Both these tests employ the principle of Wheatstone bridge for fault location.

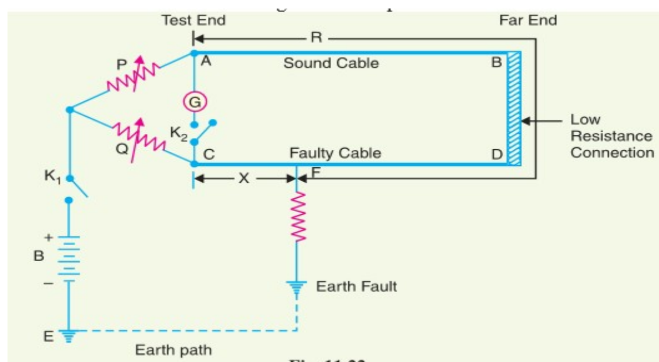
MURRAY LOOP TEST:-

The Murray loop test is the most common and accurate method of locating earth fault or short-circuit fault in underground cables.

(i) Earth fault: Fig. shows the circuit diagram for locating the earth fault by Murray loop test. Here AB is the sound cable and CD is the faulty cable; the earth fault occurring at point F.

The far end D of the faulty cable is joined to the far end B of the sound cable through a low resistance link. Two variable resistances P and Q are joined to ends A and C respectively and serve as the ratio arms of the Wheatstone bridge.

Let R = resistance of the conductor loop upto the fault from the test end X = resistance of the other length of the loop



Note that P, Q, R and X are the four arms of the Wheatstone bridge. The resistances P and Q are varied till the galvanometer indicates zero deflection.

In the balanced position of the bridge, we have,

$$\frac{P}{Q} = \frac{R}{X}$$

or $$\frac{P}{Q}+1 = \frac{R}{X}+1$$

or $$\frac{P+Q}{Q} = \frac{R+X}{X}$$

If r is the resistance of each cable, then $R+X=2r$.

$$\therefore \quad \frac{P+Q}{Q} = \frac{2r}{X}$$

or $$X = \frac{Q}{P+Q}\times 2r$$

If l is the length of each cable in metres, then resistance per metre length of cable $= \frac{r}{l}$.

$\therefore$ Distance of fault point from test end is

$$d = \frac{X}{r/l} = \frac{Q}{P+Q}\times 2r \times \frac{l}{r} = \frac{Q}{P+Q}\times 2l$$

or $$d = \frac{Q}{P+Q}\times (\text{loop length}) \text{ *metres}$$

Thus the position of the fault is located. Note that resistance of the fault is in the battery circuit and not in the bridge circuit. Therefore, fault resistance does not affect the balancing of the bridge. However, if the fault resistance is high, the sensitivity of the bridge is reduced.

(ii) Short-circuit fault: Fig. shows the circuit diagram for locating the short-circuit fault by Murray loop test. Again P, Q, R and X are the four arms of the bridge. Note that fault resistance is in the battery circuit and not in the bridge circuit. The bridge in balanced by adjusting the resistances P and Q. In the balanced position of the bridge :

$$\frac{P}{Q} = \frac{R}{X}$$

or
$$\frac{P+Q}{Q} = \frac{R+X}{X} = \frac{2r}{X}$$

$\therefore$
$$X = \frac{Q}{P+Q} \times 2r$$

or
$$X = \frac{Q}{P+Q} \times \text{(loop length) metres}$$

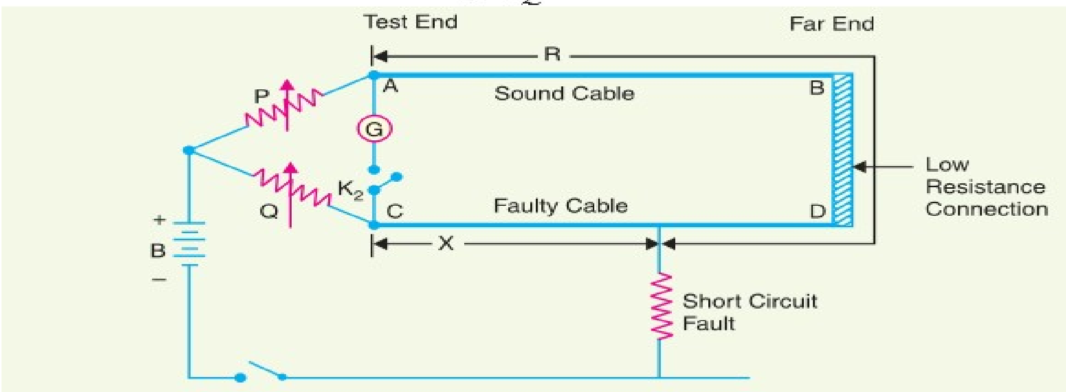


Fig. 11.23

Thus the position of the fault is located.

VARLEY LOOP TEST:-

The Varley loop test is also used to locate earth fault or short-circuit fault in underground cables. This test also employs Wheatstone bridge principle. It differs from Murray loop test in that here the ratio arms P and Q are fixed resistances. Balance is obtained by adjusting the variable resistance S connected to the test end of the faulty cable. The connection diagrams for locating the earth fault and short-circuit fault by Varley loop test are shown in Figs. 11.24 and 11.25 respectively

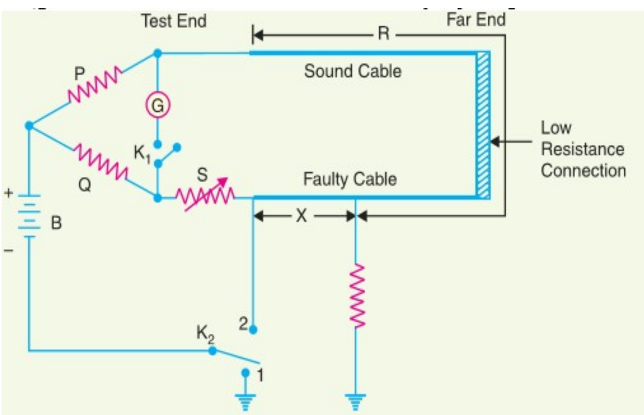


Fig. 11.24 Varley Loop Test (Earth Fault)

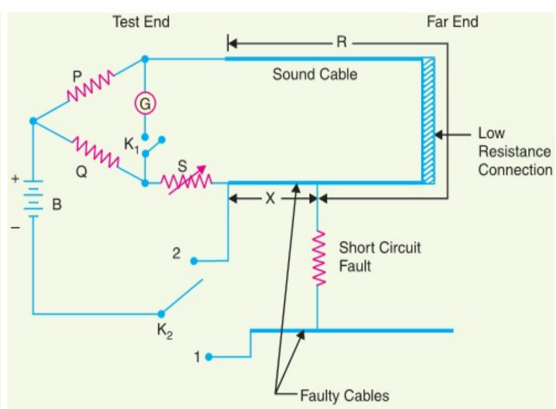


Fig. 11.25 Varley Loop Test (Short Circuit Test)

For earth fault or short-circuit fault, the key K_2 is first thrown to position 1. The variable resistance S is varied till the bridge is balanced for resistance value of S_1. Then,

$$\frac{P}{Q} = \frac{R}{X + S_1}$$

or
$$\frac{P + Q}{Q} = \frac{R + X + S_1}{X + S_1}$$

or
$$X = \frac{Q(R + X) - PS_1}{P + Q} \quad ...(i)$$

Now key K_2 is thrown to position 2 (for earth fault or short-circuit fault) and bridge is balanced with new value of resistance S_2. Then,

$$\frac{P}{Q} = \frac{R + X}{S_2}$$

or
$$(R + X)\,Q = PS_2 \quad ...(ii)$$

From eqs. (*i*) and (*ii*), we get,

$$X = \frac{P(S_2 - S_1)}{P + Q}$$

Since the values of P, Q, S_1 and S_2 are known, the value of X can be determined.

$$\text{Loop resistance} = R + X = \frac{P}{Q}S_2$$

If r is the resistance of the cable per metre length, then,

Distance of fault from the test end is

$$d = \frac{X}{r} \text{ metres}$$

Example 11.22. *In a test by Murray loop for ground fault on 500 m of cable having a resistance of 1·6 Ω/km, the faulty cable is looped with a sound cable of the same length and area of cross-section. If the ratio of the other two arms of the testing network at balance is 3 : 1, find the distance of the fault from the testing end of cables.*

Solution.

$$\frac{P}{Q} = 3 \quad \text{or} \quad \frac{P + Q}{Q} = 4$$

Distance of fault from test end is

$$d = \frac{Q}{P + Q} \times \text{loop length} = \frac{1}{4} \times (2 \times 500) = \mathbf{250\ m}$$

Example 11.23. *In a test for a fault to earth on a 500 m length of cable having a resistance of 1 Ω per 1000 m, the faulty cable is looped with a sound cable of the same length but having a resistance of 2·25 Ω per 1000 m. The resistance of the other two arms of the testing network at balance are in the ratio 2·75 : 1. Calculate the distance of the fault from the testing end of the cable.*

Solution.
$$\frac{P}{Q} = 2{\cdot}75 \quad \text{or} \quad \frac{P + Q}{Q} = 2{\cdot}75 + 1 = 3{\cdot}75$$

$$\text{Resistance of loop} = \frac{1}{1000} \times 500 + \frac{2{\cdot}25}{1000} \times 500 = 1{\cdot}625\ \Omega$$

Resistance of faulty cable from test end upto fault point is

$$X = \frac{Q}{P + Q} \times (\text{loop resistance}) = \frac{1}{3{\cdot}75} \times 1{\cdot}625 = 0{\cdot}433\ \Omega$$

Distance of fault point from the testing end is

$$d = \frac{X}{1/1000} = 0{\cdot}433 \times 1000 = \mathbf{433\ m}$$

CHAPTER-8
ECONOMIC ASPECTS

e. State and explain causes of low power factor.

The cosine of angle between voltage and current in an a.c. circuit is known as power factor.

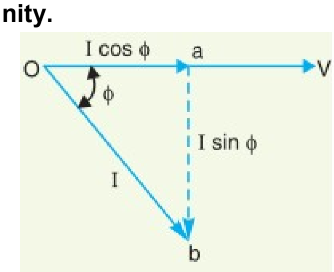
In an a.c. circuit, there is generally a phase difference φ between voltage and current. The term cos φ is called the power factor of the circuit. If the circuit is inductive, the current lags behind the voltage and the power factor is referred to as lagging. However, in a capacitive circuit, current leads the voltage and power factor is said to be leading.

Consider an inductive circuit taking a lagging current I from supply voltage V; the angleof lag being φ. The phasor diagram of the circuit is shown in Fig. The circuit current I can be resolvedinto two perpendicular components, namely

(a) I cos φ in phase with V

(b) I sin φ 90^0 out of phase with V

The component Icosφ is known as active or wattful component, whereas component Isinφ is called the reactive or wattless component. The reactive component is a measure of the power factor. If the reactive component is small, the phase angle φ is small and hence power factor cos φ will be high. Therefore, a circuit having small reactive current (i.e., I sin φ) will have high power factor and vice-versa. It may be noted that value of power factor can never be more than unity.



<u>Disadvantages of Low Power Factor</u>:-

The power factor plays an importance role in a.c. circuits since power consumed depends upon this factor.

$$P = V_L I_L \cos \varphi \qquad \text{(For single phase supply)}$$

$$\therefore \quad I_L = \frac{P}{V_L \cos \varphi} \qquad ...(i)$$

$$P = \sqrt{3}\, V_L I_L \cos \varphi \qquad \text{(For 3 phase supply)}$$

$$\therefore \quad I_L = \frac{P}{\sqrt{3}\, V_L \cos \varphi} \qquad ...(ii)$$

It is clear from above that for fixed power and voltage, the load current is inversely proportional to the power factor. Lower the power factor, higher is the load current and *vice-versa*. A power factor less than unity results in the following disadvantages :

(i) Large kVA rating of equipment. The electrical machinery (*e.g.*, alternators, transformers, switchgear) is always rated in *kVA.

Now, $$\text{kVA} = \frac{\text{kW}}{\cos \varphi}$$

It is clear that kVA rating of the equipment is inversely proportional to power factor. The smaller the power factor, the larger is the kVA rating. Therefore, at low power factor, the kVA rating of the equipment has to be made more, making the equipment larger and expensive.

(ii) Greater conductor size. To transmit or distribute a fixed amount of power at constant voltage, the conductor will have to carry more current at low power factor. This necessitates large conductor size.

(iii) Large copper losses. The large current at low power factor causes more I^2R losses in all the elements of the supply system. This results in poor efficiency.

(iv) Poor voltage regulation. The large current at low lagging power factor causes greater voltage drops in alternators, transformers, transmission lines and distributors. This results Gives poor voltage regulation.

(v) Reduced handling capacity of system. The lagging power factor reduces the handling capacity of all the elements of the system.

<u>Causes of Low Power Factor:-</u>

The following are the causes of low power factor:-

(ii) Most of the a.c. motors are of induction type (1φ and 3φ induction motors) which have low lagging power factor. These motors work at a power factor which is extremely small on light load (0·2 to 0·3) and rises to 0·8 or 0·9 at full load.

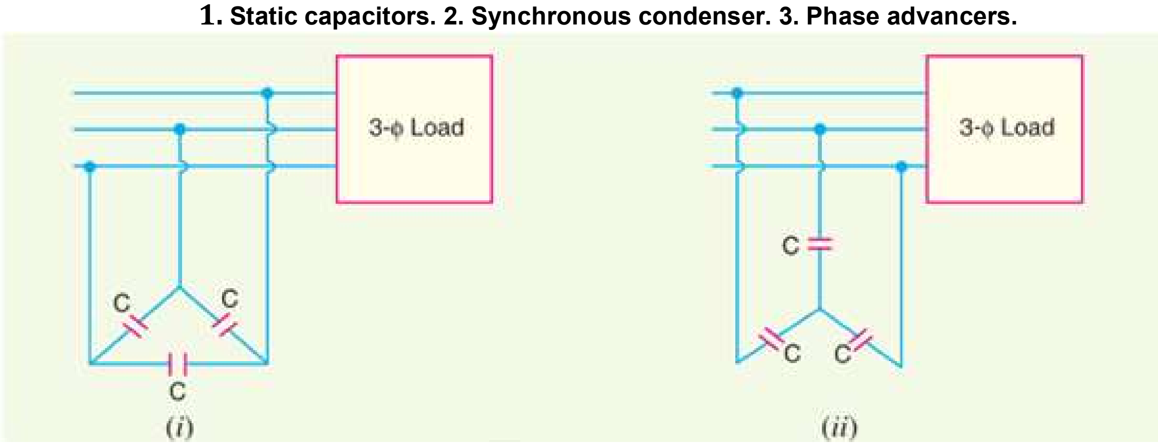
(iii) Arc lamps, electric discharge lamps and industrial heating furnaces operate at low lagging power factor.

(iv) The load on the power system is varying; being high during morning and evening and low at other times. During low load period, supply voltage is increased which increases the magnetisation current. This results in the decreased power factor.

Explain methods of improvement of power factor.

Normally, the power factor of the whole load on a large generating station is in the region of 0·8 to 0·9. However, sometimes it is lower and in such cases it is generally desirable to take special steps to improve the power factor. This can be achieved by the following equipment:

1. Static capacitors. 2. Synchronous condenser. 3. Phase advancers.



1. Static capacitor. The power factor can be improved by connecting capacitors in parallel with the equipment operating at lagging power factor. The capacitor (generally known as static

capacitor) draws a leading current and partly or completely neutralises the lagging reactive component of load current. This raises the power factor of the load. For three-phase loads, the capacitors can be connected in delta or star as shown in Fig. Static capacitors are invariably used for power factor improvement in factories.

Advantages

(i) They have low losses.

(ii) They require little maintenance as there are no rotating parts.

(iii) They can be easily installed as they are light and require no foundation.

(iv) They can work under ordinary atmospheric conditions.

Disadvantages

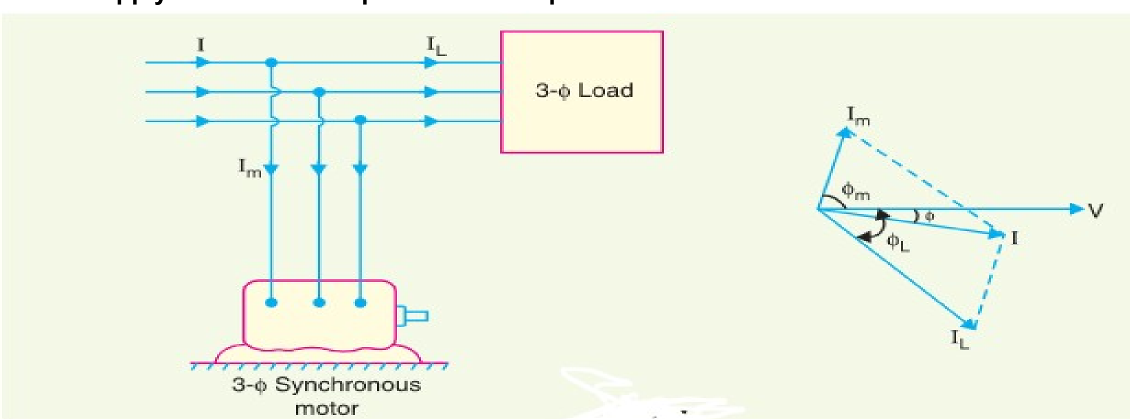
(i) They have short service life ranging from 8 to 10 years.

(ii) They are easily damaged if the voltage exceeds the rated value.

(iii) Once the capacitors are damaged, their repair is uneconomical.

2. Synchronous condenser. A synchronous motor takes a leading current when over-excited and, therefore, behaves as a capacitor. An over-excited synchronous motor running on no load is known as synchronous condenser. When such a machine is connected in parallel with the supply, it takes a leading current which partly neutralises the lagging reactive component of the load. Thus the power factor is improved.

Fig. Shows the power factor improvement by synchronous condenser method. The 3φ load takes current I_L at low lagging power factor cos φ_L. The synchronous condenser takes a current Im which leads the voltage by an angle φm. The resultant current I is the phasor sum of Im and I_L and lags behind the voltage by an angle φ. It is clear that φ is less than φ_L so that cos φ is greater than cos φ_L. Thus the power factor is increased from cos φ_L to cos φ. Synchronous condensers are generally used at major bulk supply substations for power factor improvement.



Advantages

(i) By varying the field excitation, the magnitude of current drawn by the motor can be changed by any amount. This helps in achieving stepless control of power factor.

(ii) The motor windings have high thermal stability to short circuit currents.

(iii) The faults can be removed easily.

Disadvantages

(i) There are considerable losses in the motor.

(ii) The maintenance cost is high.

(iii) It produces noise.

(iv) Except in sizes above 500 kVA, the cost is greater than that of static capacitors of the same rating.

(v) As a synchronous motor has no self-starting torque, therefore, auxiliary equipment has to be provided for this purpose.

[Note. The reactive power taken by a synchronous motor depends upon two factors, the d.c. field excitation and the mechanical load delivered by the motor. Maximum leading power is taken by a synchronous motor with maximum excitation and zero loads.]

3. Phase advancers. Phase advancers are used to improve the power factor of induction motors. The low power factor of an induction motor is due to the fact that its stator winding draws exciting current which lags behind the supply voltage by 90o. If the exciting ampere turns can be provided from some other a.c. source, then the stator winding will be relieved of exciting current and the power factor of the motor can be improved. The phase advancer is mounted on the same shaft as the main motor and is connected in the rotor circuit of the motor.

Phase advancers have two principal advantages. Firstly, as the exciting ampere turns are supplied at slip frequency, therefore, lagging kVAR drawn by the motor are considerably reduced. Secondly, phase advancer can be conveniently used where the use of synchronous motors is inadmissible.

Define & explain Load curves.

VARIABLE LOAD ON POWER STATION:-

The load on a power station varies from time to time due to uncertain demands of the consumers and is known as variable load on the station.

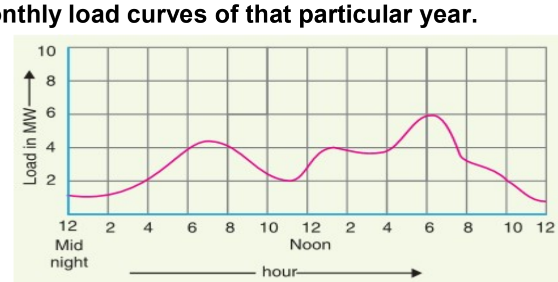
LOAD CURVE:-

The curve showing the variation of load on the power station with respect to (w.r.t) time is known as a load curve.

The load on a power station is never constant; it varies from time to time. These load variations during the whole day (i.e., 24 hours) are recorded half-hourly or hourly and are plotted against time on the graph. The curve thus obtained is known as daily load curve.

The monthly load curve can be obtained from the daily load curves of that month.

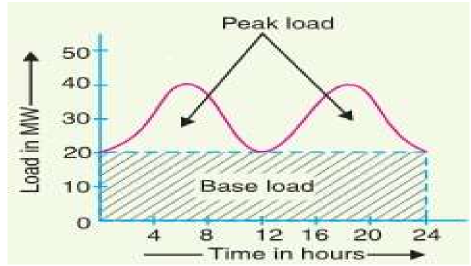
The yearly load curve is obtained by considering the monthly load curves of that particular year.



BASE LOAD AND PEAK LOAD ON POWER STATION

Base load. The unvarying load which occurs almost the whole day on the station is known as base load.

Referring to the load curve Fig. it is clear that 20 MW of load has to be supplied by the station at all times of day and night i.e. throughout 24 hours. Therefore, 20 MW is the base load of the station.



Peak load. The various peak demands of load over and above the base load of the station is known aspeak load.

It is clear that there are peak demands of load excluding base load.

Define & explain Demand factor.

Demand factor: "It is the ratio of maximum demand on the power station to its connected load" i.e.,

$$\text{Demand factor} = \frac{\text{Maximum demand}}{\text{Connected load}}$$

The value of demand factor is usually less than 1.

It is expected because maximum demand on the power station is generally less than the connected load. If the maximum demand on the power station is 80 MW and the connected load is 100 MW, then demand factor = 80/100 = 0·8. The knowledge of demand factor is vital in determining the capacity ofthe plant equipment.

Define & explain Maximum demand.

Maximum demand: "It is the greatest demand of load on the power station during a given period."

The load on the power station varies from time to time. The maximum of all the demands that have occurred during a given period (say a day) is the maximum demand. Thus referring back to the load curve of Fig. 3.2, the maximum demand on the power station during the day is 6 MW and it occursat 6 P.M. Maximum demand is generally less than the connected load because all the consumers do not switch on their connected load to the system at a time. The knowledge of maximum demand is very important as it helps in determining the installed capacity of the station. The station must be capable of meeting the maximum demand.

Connected load: "It is the sum of continuous ratings of all the equipment connected to supply system."

A power station supplies load to thousands of consumers. Each consumer has certain equipment installed in his premises. The sum of the continuous ratings of all the equipment in the consumer's premises is the "connected load" of the consumer. For instance, if a consumer has connections of five 100-watt lamps and a power point of 500 watts, then connected load of the consumer is 5 × 100 + 500
= 1000 watts. The sum of the connected loads of all the consumers is the connected load to the power station.

Define & explain Load factor.

Average load. The average of loads occurring on the power station in a given period (day or month or year) is known as average load or average demand.

$$\text{Daily average load} = \frac{\text{No. of units (kWh) generated in a day}}{24 \text{ hours}}$$

$$\text{Monthly average load} = \frac{\text{No. of units (kWh) generated in a month}}{\text{Number of hours in a month}}$$

$$\text{Yearly average load} = \frac{\text{No. of units (kWh) generated in a year}}{8760 \text{ hours}}$$

Load factor: "The ratio of average load to the maximum demand during a given period is known as load factor" i.e.,

$$\text{Load factor} = \frac{\text{Average load}}{\text{Max. demand}}$$

If the plant is in operation for T hours,

$$\text{Load factor} = \frac{\text{Average load} \times \text{T}}{\text{Max. demand} \times \text{T}}$$

$$= \frac{\text{Units generated in T hours}}{\text{Max. demand} \times \text{T hours}}$$

The load factor may be daily load factor, monthly load factor or annual load factor if the time period considered is a day or month or year. Load factor is always less than 1 because average load is smaller than the maximum demand. The load factor plays key role in determining the overall cost per unit generated. Higher the load factor of the power station, lesser* will be the cost per unit generated.

$$\text{Diversity factor} = \frac{\text{Sum of individual max. demands}}{\text{Max. demand on power station}}$$

Define & explain Diversity factor

Diversity factor: "The ratio of the sum of individual maximum demands to the maximum demand on power station is known as diversity factor" i.e.,

$$\text{Plant capacity factor} = \frac{\text{Actual energy produced}}{\text{Max. energy that could have been produced}}$$

$$= \frac{\text{Average demand} \times \text{T}^{**}}{\text{Plant capacity} \times \text{T}}$$

$$= \frac{\text{Average demand}}{\text{Plant capacity}}$$

A power station supplies load to various types of consumers whose maximum demands generally do not occur at the same time. Therefore, the maximum demand on the power station is always less than the sum of individual maximum demands of the consumers. Obviously, diversity factor will always be greater than 1. The greater the diversity factor, the lesser is the cost of generation of power.

Define & explain Plant capacity factor.

Plant capacity factor: "It is the ratio of actual energy produced to the maximum possible energy that could have been produced during a given period" i.e.,

Thus if the considered period is one year,

$$\text{Annual plant capacity factor} = \frac{\text{Annual kWh output}}{\text{Plant capacity} \times 8760}$$

The plant capacity factor is an indication of the reserve capacity of the plant.

A power station is so designed that it has some reserve capacity for meeting the increased load demand in future. Therefore, the installed capacity of the plant is always somewhat greater than the maximum demand on the plant.

Reserve capacity = Plant capacity − Max. demand

It is interesting to note that difference between load factor and plant capacity factor is an indication of reserve capacity. If the maximum demand on the plant is equal to the plant capacity, then load factor and plant capacity factor will have the same value. In such a case, the plant will have no reserve capacity.

Define & explain peak load and Base load on power station

The changing load on the power station makes its load curve of variable nature. Fig. 3.13. Shows the typical load curve of a power station. It is clear that load on the power station varies from time to time. However, a close look at the load curve reveals that load on the power station can be considered in two parts, namely;

(i) Base load

(ii) Peak load

Base load: The unvarying load which occurs almost the whole day on the station is known as baseload

Referring to the load curve of Fig. 3.13, it is clear that 20 MW of load has to be supplied by the station at all times of day and night i.e. throughout 24 hours. Therefore, 20 MW is the base load ofthe station.

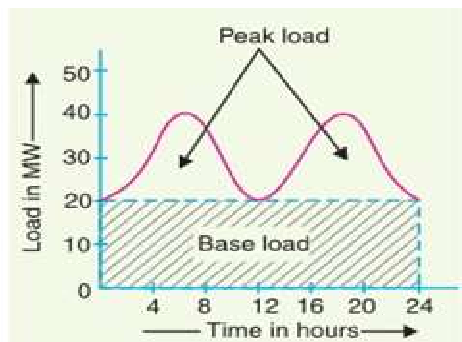


Fig. 3.13

Peak load: The various peak demands of load over and above the base load of the station is known as peak load.

Referring to the load curve of Fig. 3.13, it is clear that there are peak demands of load excluding base load. These peak demands of the station generally form a small part of the total load and may occur throughout the day

CHAPTER-9

TYPES OF TARIFF

9.1 Explain flat rate and two part tariff and block rate tariff with problems

The rate at which electrical energy is supplied to a consumer is known as tariff. Tariff should include the total cost of producing and supplying electrical energy plus the profit.

Objectives of tariff. Like other commodities, electrical energy is also sold at such a rate so that it not only returns the cost but also earns reasonable profit. Therefore, a tariff should include the following items:

(i) Recovery of cost of producing electrical energy at the power station.

(ii) Recovery of cost on the capital investment in transmission and distribution systems.

(iii) Recovery of cost of operation and maintenance of supply of electrical energy e.g., metering equipment, billing etc.

(iv) A suitable profit on the capital investment.

CHARACTERISTICS OF TARIFF:-

A tariff must have the following desirable characteristics:

(i) Proper return : The tariff should be such that it ensures the proper return from each consumer.

(ii) Fairness : The tariff must be fair so that different types of consumers are satisfied with the rate of charge of electrical energy.

(iii) Simplicity : The tariff should be simple so that an ordinary consumer can easily understand it.

(iv) Reasonable profit : The profit element in the tariff should be reasonable.

(v) Attractive : The tariff should be attractive so that a large number of consumers are encouraged to use electrical energy.

TYPES OF TARIFF:-

There are several types of tariff like:

1. Simple tariff. When there is a fixed rate per unit of energy consumed, it is called a simple tariff or uniform rate tariff.

In this type of tariff, the price charged per unit is constant i.e., it does not vary with increase or decrease in number of units consumed. The consumption of electrical energy at the consumer's terminals is recorded by means of an energy meter. This is the simplest of all tariffs and is readily understood by the consumers.

Disadvantages

(i) There is no discrimination between different types of consumers since every consumer has to pay equitably for the fixed* charges.

(ii) The cost per unit delivered is high.

(iii) It does not encourage the use of electricity.

2. Flat rate tariff. When different types of consumers are charged at different uniform per unit rates, it is called a flat rate tariff. In this type of tariff, the consumers are grouped into different classes and each class of consumers is charged at a different uniform rate. For instance, the flat rate per kWh for lighting load may be 60 paise, whereas it may be slightly less† (say 55 paise per kWh) for power load.

Disadvantages

(i) Since the flat rate tariff varies according to the way the supply is used, separate meters are required for lighting load, power load etc.

(ii) A particular class of consumers is charged at the same rate irrespective of the magnitude of energy consumed.

3. Block rate tariff. When a given block of energy is charged at a specified rate and the succeeding blocks of energy are charged at progressively reduced rates, it is called a block rate tariff.

In block rate tariff, the energy consumption is divided into blocks and the price per unit is fixed in each block. The price per unit in the first block is the highest and it is progressively reduced for the succeeding blocks of energy.

The advantage of such a tariff is that the consumer gets an incentive to consume more electrical energy. This increases the load factor of the system and hence the cost of generation is reduced.

4. two-part tariff. When the rate of electrical energy is charged on the basis of maximum demand of the consumer and the units consumed, it is called a two-part tariff.

In two-part tariff, the total charge to be made from the consumer is split into two components viz., fixed charges and running charges. The fixed charges depend upon the maximum demand of the consumer while the running charges depend upon the number of units consumed by the consumer. Thus,the consumer is charged at a certain amount per kW of maximum demand plus a certain amount per kWhof energy consumed.

$$\text{Total charges} = \text{Rs}\,(b \times \text{kW} + c \times \text{kWh})$$

where, b = charge per kW of maximum demand

c = charge per kWh of energy consumed

This type of tariff is mostly applicable to industrial consumers who have appreciable maximum demand.

Advantages

(i) It is easily understood by the consumers.

(ii) It recovers the fixed charges which depend upon the maximum demand of the consumer but are independent of the units consumed.

Disadvantages

(i) The consumer has to pay the fixed charges irrespective of the fact whether he has consumed or not consumed the electrical energy.

(ii) There is always error in assessing the maximum demand of the consumer.

5. Maximum demand tariff. It is similar to two-part tariff with the only difference that the maximum demand is actually measured by installing maximum demand meter in the premises of the consumer.

6. Power factor tariff. The tariff in which power factor of the consumer's load is taken intoconsideration is known as power factor tariff. In an a.c. system, power factor plays an important role. A low* power factor increases the rating of station equipment and line losses. Therefore, a consumer having low power factor must be penalized.

CHAPTER-10
SUBSTATION

Draw and explain layout of LT. HT and EHT substation.

The electric power is produced at the power stations and It is delivered to the consumers through a large network of transmission and distribution. At many places in the line of the power system, it may be desirable and necessary to change some characteristic (e.g. voltage, a.c. to d.c., frequency, p.f. etc.) of electric supply. This is accomplished by suitable apparatus called substation.

For example, generation voltage (11 kV or 6·6 kV) at the power station is stepped up to high voltage (say 220 kV or 132 kV) for transmission of electric power. The assembly of apparatus (e.g. transformer etc.) used for this purpose is the sub-station.

"The assembly of apparatus used to change some characteristic (e.g. voltage, a.c. to d.c., frequency, p.f. etc.) of electric supply is called a sub-station."

The following are the important points which must be kept in view while laying out a sub-station:

(i) It should be located at a proper site. As far as possible, it should be located at the centre of gravity of load.

(ii) It should provide safe and reliable arrangement. For safety, consideration must be given to the maintenance of regulation clearances, facilities for carrying out repairs and maintenance, abnormal occurrences such as possibility of explosion or fire etc. For reliability, consideration must be given for good design and construction, the provision of suitable protective gear etc.

(iii) It should be easily operated and maintained.

(iv) It should involve minimum capital cost.

<u>CLASSIFICATION OF SUB-STATION:</u>

1. According to service requirement. A sub-station may be called upon to change voltage levelor improve power factor or convert a.c. power into d.c. power etc. According to the service requirement, sub-stations may be classified into:

(i) Transformer sub-stations. Those sub-stations which change the voltage level of electric supply are called transformer sub-stations. These sub-stations receive power at some voltage and deliver it at some other voltage. Obviously, transformer will be the main component in such substations. Most of the sub-stations in the power system are of this type.

(ii) Switching sub-stations. These sub-stations do not change the voltage level i.e. incoming and outgoing lines have the same voltage. However, they simply perform the switching operations of power lines.

(iii) Power factor correction sub-stations. Those sub-stations which improve the power factorof the system are called power factor correction sub-stations. Such sub-stations are generally located at the receiving end of transmission lines. These sub-stations generally use synchronous condensers as thepower factor improvement equipment.

(iv) Frequency changer sub-stations. Those sub-stations which change the supply frequency are known as frequency changer sub-stations. Such a frequency change may be required for industrial utilisation.

(v) **Converting sub-stations.** Those sub-stations which change a.c. power into d.c. power are called converting sub-stations. These sub-stations receive a.c. power and convert it into d.c. power with suitable apparatus (e.g. ignitron) to supply for such purposes as traction, electroplating, electric welding etc.

(vi) **Industrial sub-stations.** Those sub-stations which supply power to individual industrial concerns are known as industrial sub-stations.

2. **According to constructional features.** A sub-station has many components (e.g. circuit breakers, switches, fuses, instruments etc.) which must be housed properly to ensure continuous and reliable service. According to constructional features, the sub-stations are classified as :

(i) **Indoor sub-stations.** For voltages upto 11 kV, the equipment of the sub-station is installed indoor because of economic considerations. However, when the atmosphere is contaminated with impurities, these sub-stations can be erected for voltages upto 66 kV.

(ii) **Outdoor sub-stations.** For voltages beyond 66 kV, equipment is invariably installed outdoor. It is because for such voltages, the clearances between conductors and the space required for switches, circuit breakers and other equipment becomes so great that it is not economical to install the equipment indoor.

(iii) **Underground sub-stations.** In thickly populated areas, the space available for equipment and building is limited and the cost of land is high. Under such situations, the sub-station is created underground. The reader may find further discussion on underground sub-stations.

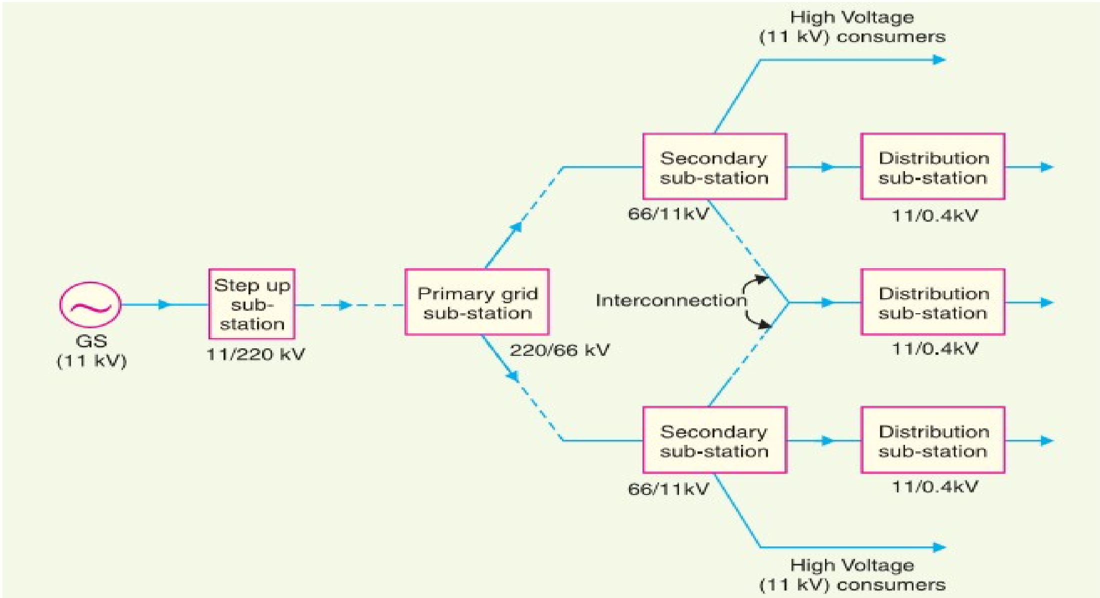
(iv) **Pole-mounted sub-stations.** This is an outdoor sub-station with equipment installed overhead on H-pole or 4-pole structure. It is the cheapest form of sub-station for voltages not exceeding 11kV (or 33 kV in some cases). Electric power is almost distributed in localities through such substations.

TRANSFORMER SUB-STATION:

The majority of the sub-stations in the power system are concerned with the changing of voltage level of electric supply. These are known as transformer sub-stations because transformer is the main component employed to change the voltage level. Depending upon the purpose served, transformer sub-stations may be classified into:

(i) Step-up sub-station	(ii) Primary grid sub-station
(iii) Secondary sub-station	(iv) Distribution sub-station

Fig. shows the block diagram of a typical electric supply system indicating the position of above types of sub-stations. It may be noted that it is not necessary that all electric supply schemes include all the stages shown in the figure. For example, in a certain supply scheme there may not be secondary sub-stations and in another case, the scheme may be so small that there are only distribution sub- stations.



(i) Step-up sub-station. The generation voltage (11 kV in this case) is stepped up to high voltage (220kV) to affect economy in transmission of electric power. The sub-stations which accomplish this job are called step-up sub-stations. These are generally located in the power houses and are of outdoor type.

(ii) Primary grid sub-station. From the step-up sub-station, electric power at 220 kV is transmitted by3-phase, 3-wire overhead system to the outskirts of the city. Here, electric power is received by the primary grid sub-station which reduces the voltage level to 66 kV for secondary transmission. The primary grid sub-station is generally of outdoor type.

(iii) Secondary sub-station. From the primary grid sub-station, electric power is transmitted at 66 kV by 3-phase, 3-wire system to various secondary sub-stations located at the strategic points in the city. Ata secondary sub-station, the voltage is further stepped down to 11 kV. The 11 kV lines run along the important road sides of the city. It may be noted that big consumers (having demand more than 50 kW) are generally supplied power at 11 kV for further handling with their own substations. The secondary sub-stations are also generally of outdoor type.

(iv) Distribution sub-station. The electric power from 11 kV lines is delivered to distribution sub-stations. These sub-stations are located near the consumers localities and step down the voltage to 400 V, 3-phase, 4-wire for supplying to the consumers. The voltage between any two phases is 400V and between any phase and neutral it is 230 V. The single phase residential lighting load is connected between any one phase and neutral whereas 3-phase, 400V motor load is connected across 3-phase lines directly. It may be worthwhile to mention here that majority of the distribution substations are of pole-mounted type.

KEY DIAGRAM OF 66/11KV SUBSTATION

(i) There are two 66 kV incoming lines marked 'incoming 1' and 'incoming 2' connected to the bus-bars. Such an arrangement of two incoming lines is called a double circuit. Each incoming line is capable of supplying the rated sub-station load. Both these lines can be loaded simultaneously to share the sub-station load and any one line can be called upon to meet the entire load. The double circuit arrangement increases the reliability of the system. In case there is a breakdown of one incoming line, the continuity of supply can be maintained by the other line.

(ii) The sub-station has duplicate bus-bar system; one 'main bus-bar' and the other spare busbar. The incoming lines can be connected to either bus-bar with the help of a bus-coupler which consists of a circuit breaker and isolators. The advantage of double bus-bar system is that if repair is to

be carried on one bus-bar, the supply need not be interrupted as the entire load can be transferred to the other bus.

(iii) There is an arrangement in the sub-station by which the same 66 kV double circuit supply is going out i.e. 66 kV double circuit supply is passing through the sub-station. The outgoing 66 kV double circuit line can be made to act as incoming line.

(iv) There is also an arrangement to step down the incoming 66 kV supply to 11 kV by two unitsof 3-phase transformers; each transformer supplying to a separate bus-bar. Generally, one transformer supplies the entire sub-station load while the other transformer acts as a standby unit. If need arises, both the transformers can be called upon to share the sub-station load. The 11 kV outgoing lines feed to the distribution sub-stations located near consumers localities.

(v) Both incoming and outgoing lines are connected through circuit breakers having isolators on their either end. Whenever repair is to be carried over the line towers, the line is first switched off and then earthed.

(vi) The potential transformers (P.T.) and current transformers (C.T.) and suitably located for supply to metering and indicating instruments and relay circuits (not shown in the figure). The P.T. is connected right on the point where the line is terminated. The CTs are connected at the terminals of each circuit breaker.

(vii) The lightning arresters are connected near the transformer terminals (on H.T. side) to protect them from lightning strokes.

(viii) There are other auxiliary components in the sub-station such as capacitor bank for power factor improvement, earth connections, local supply connections, d.c. supply connections etc. However, these have been omitted in the key diagram for the sake of simplicity.

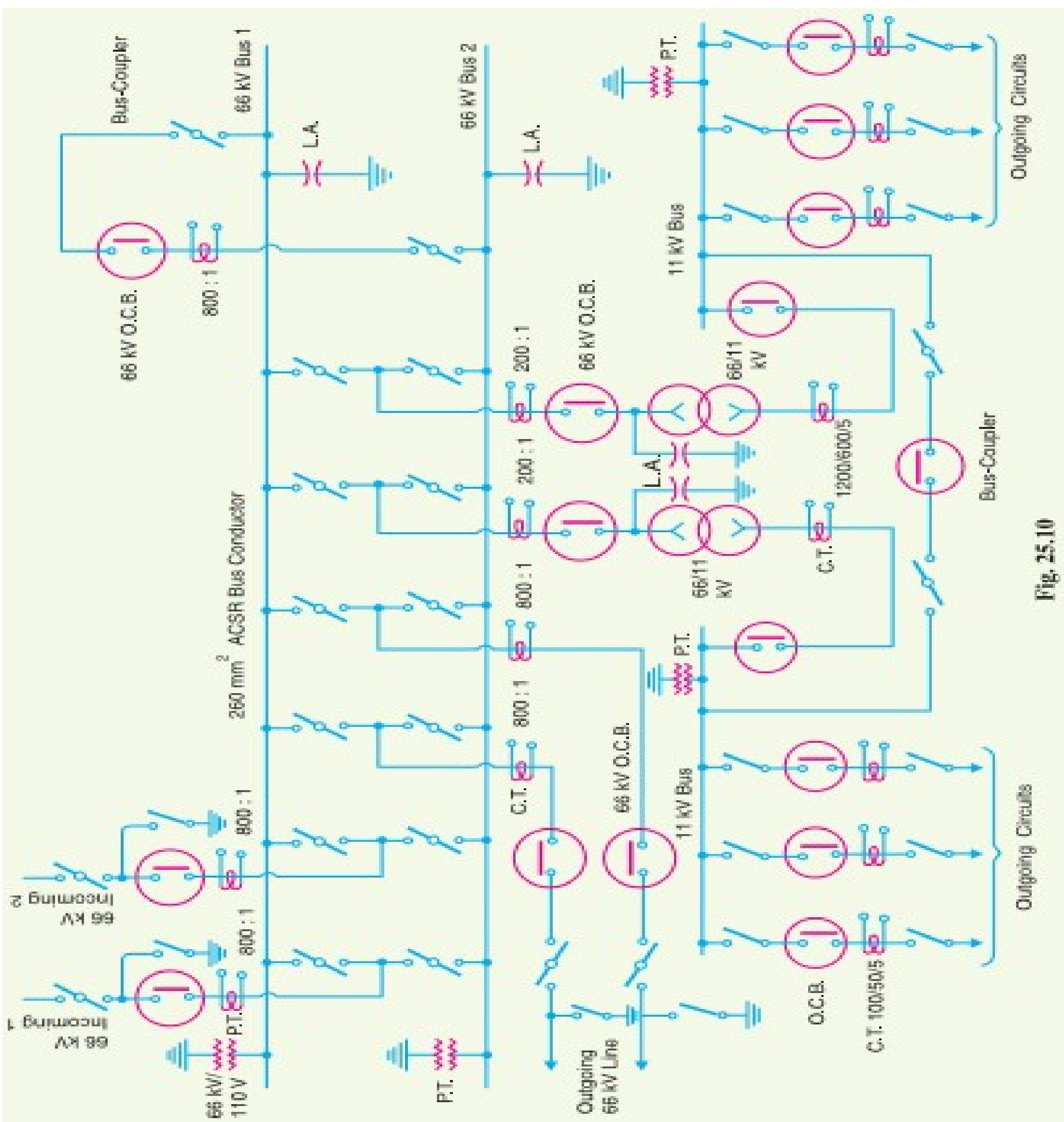


Fig. 25.10

KEY DIAGRAM OF A 11 KV/400 V INDOOR SUB-STATION (LT SUBSTATION)

The key diagram of this sub-station can be explained as under :

(i) The 3-phase, 3-wire 11 kV line is tapped and brought to the gang operating switch installed near the sub-station. The G.O. switch consists of isolators connected in each phase of the 3phase line.

(ii) From the G.O. switch, the 11 kV line is brought to the indoor sub-station as underground cable. It is fed to the H.T. side of the transformer (11 kV/400 V) via the 11 kV O.C.B. The transformersteps down the voltage to 400 V, 3-phase, 4-wire.

(iii) The secondary of transformer supplies to the bus-bars via the main O.C.B. From the busbars, 400 V, 3-phase, 4-wire supply is given to the various consumers via 400 V O.C.B. The voltage between any two phases is 400 V and between any phase and neutral it is 230 V. The single phase residential load is connected between any one phase and neutral whereas 3phase, 400 V motor load is connected across 3-phase lines directly.

(iv) The CTs are located at suitable places in the sub-station circuit and supply for the metering and indicating instruments and relay circuits.

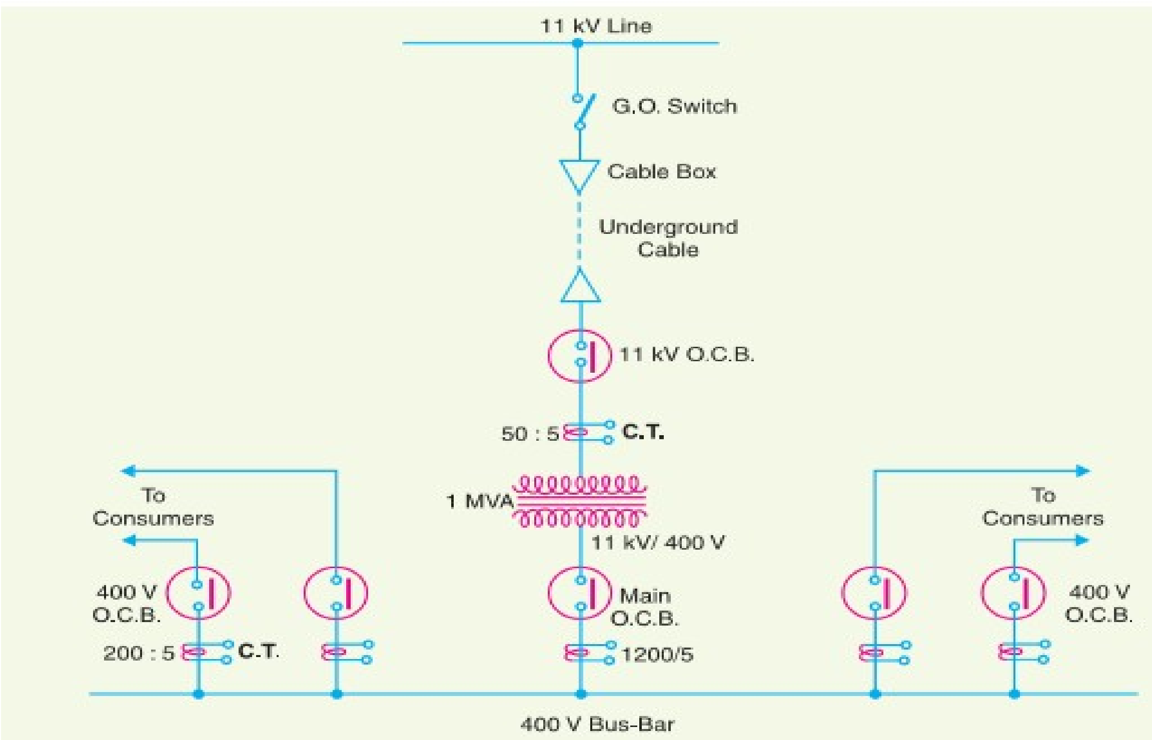
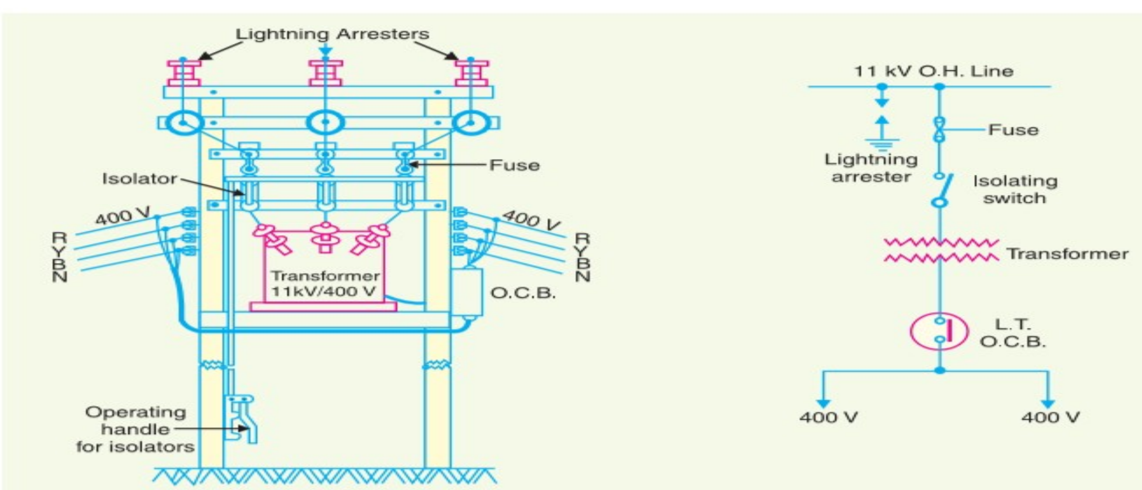


Fig. 25.11



(i) (ii)

9.2 Draw and Explain Earthing of Substation, transmission and distribution lines.

In power system, grounding or earthing means connecting frame of electrical equipment (non-current carrying part) or some electrical part of the system (e.g. neutral point in a star-connected system, one conductor of the secondary of a transformer etc.) to earth i.e. soil. This connection to earth may be through a conductor or some other circuit element (e.g. a resistor, a circuit breaker etc.) depending upon the situation.

The process of connecting the metallic frame (i.e. non-current carrying part) of electrical equipment or some electrical part of the system (e.g. neutral point in a star-connected system, one conductor of the secondary of a transformer etc.) to earth (i.e. soil) is called grounding or earthing.

It is strange but true that grounding of electrical systems is less understood aspect of power system. Nevertheless, it is a very important subject. If grounding is done systematically in the line of the power system, we can effectively prevent accidents and damage to the equipment of the power system and at the same time continuity of supply can be maintained. Grounding or earthing may be classified as :

(i) Equipment grounding

(ii) System grounding.

Equipment grounding deals with earthing the non-current-carrying metal parts of the electrical equipment. On the other hand, system grounding means earthing some part of the electrical system e.g. earthing of neutral point of star-connected system in generating stations and sub-stations.

The process of connecting non-current-carrying metal parts (i.e. metallic enclosure) of the electrical equipment to earth (i.e. soil) in such a way that in case of insulation failure, the enclosure effectively remains at earth potential is called equipment grounding.

The process of connecting some electrical part of the power system (e.g. neutral point of a star connected system, one conductor of the secondary of a transformer etc.) to earth (i.e. soil) is called system grounding.

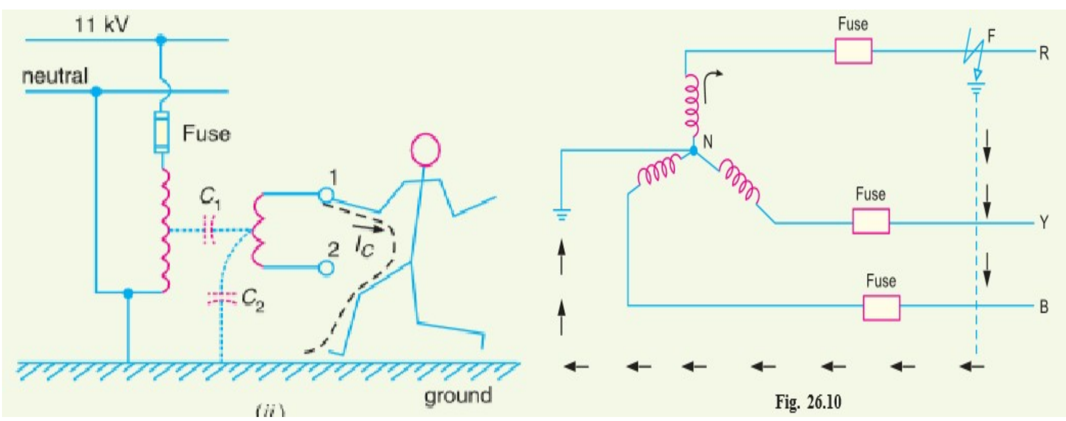


Fig. 26.10

The modern high-voltage 3-phase systems employ grounded neutral owing to a number of advantages.

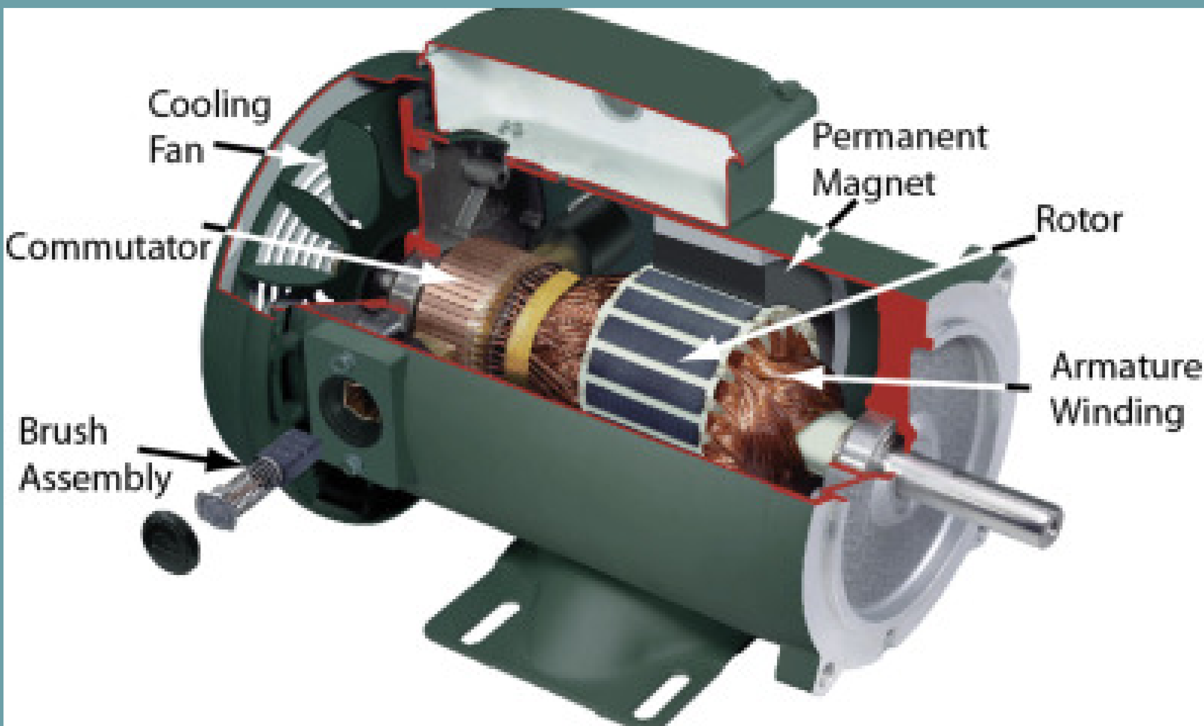
The process of connecting neutral point of 3-phase system to earth (i.e. soil) either directly or through some circuit element (e.g. resistance, reactance etc.) is called neutral grounding.

Neutral grounding provides protection to personal and equipment. It is because during earth fault, the current path is completed through the earthed neutral and the protective devices (e.g. a fuse etc.) operate to isolate the faulty conductor from the rest of the system. This point is illustrated in Fig. 26.10.

GANDHI ACADEMY OF TECHNOLOGY AND ENGINEERING

ENERGY CONVERSION II

5Th Sem Electrical
(As per SCTE&VT Syllabus)



Prepared By :
Mahesh Kumar Mishra
Dept. Of Electrical Engg.

CHAPTER-I

Induction Motor

Construction :

The induction motor mainly divided in to two parts.

(1) Stator (2) Rotor

In case of D. C. Motor basically it is divided into two main parts (i) Yoke (ii) Armature. Yoke is outer & stationary part, similarly the outer portion of the induction motor is known as stator. It is also stationary part of the induction motor. The stator of the induction motor is cylindrical in shape.

The inner part of D. C. Motor i.e., armature is rotating in nature. Similarly the rotating part of the induction motor is known as rotor. The rotor lies inside the stator. It is cylindrical in shape.

Rotor is divided into two types.

(i) Squirrel cage Rotor

(ii) Phase wound Rotor or Slip ring Rotor,

Figure shows the disassembled view of an induction motor with squirrel cage rotor.

(a) Stator (b) Rotor (c) bearing shields (d) Fan (e) Ventilation grill (f) terminal box.

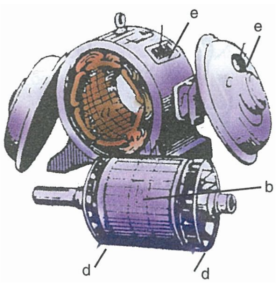


Fig 1.1

Similarly figure shows the disassembled view of a slip ring motor (a) stator (b) rotor (c) bearing shields (d) Fan (e) Ventilation grill (f) Terminal box (g) Slip ring (h) brushes & brush holder.

Production of Rotating Magnetic Field :

When 3 – phase stationary coils are fed with 3 – phase supply, a uniformly rotating magnetic flux of constant magnitude will produce.

It will now be shown that when three – phase winding displaced in space by 120^0, are fed by three phase currents, displaced in time by 120^0, they produce a resultant magnetic flux, which rotates in space as if actual magnetic poles were being rotated mechanically.

The principle of a 3 – phase, two pole stator having three identical windings placed 120^0 space degree apart as shown in fig – 1.2. The flux due to three phase windings is shown in fig 1.3.

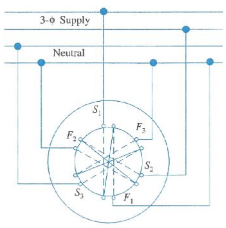


Fig 1.2

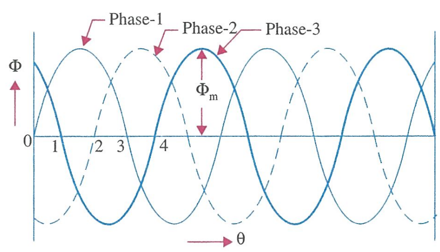


Fig 1.3

Let the maximum value of flux due to any one of the three phases be ϕ_m. The resultant flux ϕ_r, at any instant is given by the vector sum of the individual fluxes ϕ_1, ϕ_2 and ϕ_3 due to three phases. Considering values of ϕ_r at four instants i.e. $1/6^{th}$ time period apart corresponding to points marked 0, 1, 2 & 3.

Proof :

Case – 1 : Resultant flux at origin i.e. when $\theta = 0^0$ At that time $\phi_1 = 0$,

$\phi_2 = \phi m \text{ Sin} < -120^0 = -\dfrac{\sqrt{3}}{2}\ \phi m$ $\phi_3 = \phi m \text{ Sin} < -240^0 = -\dfrac{\sqrt{3}}{2}\ \phi m.$

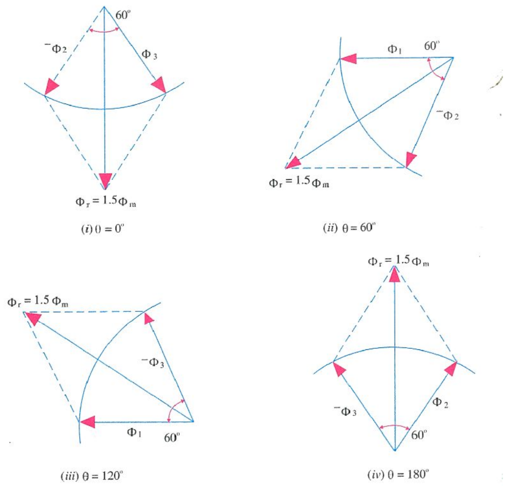


Fig 1.4

Resultant flux ϕ_r :

As per law of parallerogram

$$\phi_r^2 = \phi_2^2 + \phi_3^2 + 2\,\phi_2 \cdot \phi_3 \cdot \cos 60^0$$

$$\Rightarrow \phi_r^2 = \left(\frac{\sqrt{3}}{2}\phi_m\right)^2 + \left(\frac{\sqrt{3}}{2}\phi_m\right)^2 + 2.\ \frac{\sqrt{3}}{2}\ \phi_m \cdot \frac{\sqrt{3}}{2}\ \phi_m\ \frac{1}{2}$$

$$\Rightarrow \phi_r^2 = \frac{3}{4}\phi_m^2 + \frac{3}{4}\phi_m^2 + \frac{3}{4}\phi_m^2$$

$$\Rightarrow \phi_r^2 = \frac{9}{4}\phi_m^2$$

$$\Rightarrow \phi_r = \frac{3}{2}\phi_m$$

$$\Rightarrow \phi_r = 1.5\ \phi_m$$

Case – II : When $\theta = 60^0$

Therefore $\phi_1 = \phi_m \text{ Sin} < 60^0 = \dfrac{\sqrt{3}}{2}\ \phi_m$

$\phi_2 = \phi_m \text{ Sin} < -120^0 + 60^0 = \phi_m \text{ Sin} < -60^0 = \dfrac{-\sqrt{3}}{2}\ \phi_m$

and $\phi_3 = \phi_m \text{ Sin} < -240^0 + 60^0 = \phi_m \text{ Sin} < -180 = 0$

case – III When $\theta = 120^0$

$\phi_1 = \phi_m \text{ Sin} < 120^0 = \dfrac{\sqrt{3}}{2}\ \phi_m$

$\phi_2 = \phi_m \text{ Sin} < -120^0 + 120^0 = \phi_m \text{ Sin} < 0^0 = 0$

$\phi_3 = \phi_m \text{ Sin} < -240^0 + 120^0 = \phi_m \text{ Sin} < -120^0 = \dfrac{-\sqrt{3}}{2}\ \phi_m$

ϕ_r can be calculated as earlier

Similarly $\phi_r = 1.5\phi_m$

Case – IV When $\theta = 180^0$

$\phi_1 = \phi_m \text{ Sin} < 180^0 = 0$

$\phi_2 = \phi_m \text{ Sin} < -120^0 + 180^0 = \phi_m \text{ Sin} < 60^0 = \dfrac{\sqrt{3}}{2}\ \phi_m$

$\phi_3 = \phi_m \text{ Sin} < -240^0 + 180^0 = \phi_m \text{ Sin} < -60^0 = \dfrac{-\sqrt{3}}{2}\ \phi_m$

Similarly ϕ_r can be calculated as earlier $\phi_r = 1.5\ \phi_m$

Hence from the above four cases we can draw a conclusion that the resultant flux (ϕ_r) inside the stator winding at any time = 1.5 ϕm and the resultant flux (ϕ_r) rotates around the stator at syncronous speed.

How the rotor rotates :

The rotor lies inside the stator. There is an air gap in between the stator and rotor. The stator slots are provided with three Phase winding.

When three phase stator windings are fed by a 3-phase supply then a rotating magnetic flux of constant magnitude will produce.

This rotating flux passes through air gap and cuts the stationary conductors on the rotor . There is also a 3-phase rotor winding on the rotor. The stator and rotor windings act as primary and secondary windings of a 3-phase transformer. The air gap acts as core of the transformer. The fluxes passes from stator to rotor winding through induction principle.

The rotating flux produces an emf in the rotor winding. The rotor winding is closed circuit. Hence current will flow in the rotor conductors. When current will flow it will produce the flux in the air gap. The flux in the rotor winding interacts with the flux in the stator winding there by producing a torque, which is responsible for the rotation of the rotor.

Slip(s) :

The rotor never succeeds in catching up with the stator field. It is really did so, then there would be no relative speed between the two, hence no rotor emf, no rotor current and so no torque to maintain rotation. That in why the rotor runs at a speed which is always less than the speed of the stator field.

The difference between synchronous speed N_s to the actual speed of the totor N_r is known as slip speed.

Slip speed = $N_s - N_r$.

Slip (s) or % of Slip (s) = $\frac{N_s - N_r}{N_s} \times 100$

$$\Rightarrow S = \frac{N_s - N_r}{N_s}$$

$\Rightarrow N_s - N_r = SN_s$

$\Rightarrow N_s - SN_S = N_r$

$\Rightarrow N_s(1\text{-}S) = N_r$

Therefore Rotor speed $N_r = N_s\,(1\text{-}S)$

Frequency of Rotor Current :

When the rotor is stationary, the frequency of rotor current is the same as the supply frequency. But when the rotor starts revolving, then the frequency depends upon the relative speed. Let the frequency of the rotor current be f'.

Hence $\quad N_s - N_r == \frac{120 f'}{P}$

As Ns = $= \dfrac{120f}{P}$

$$\Rightarrow \frac{N_s - N_r}{N_s} = \frac{120f'}{P} \times \frac{P}{120f}$$

$$\Rightarrow S = \frac{f'}{f}$$

Therefore $f' = Sf$

Hence Rotor frequency = slip x supply frequency

Torque of an Induction Motor :

The torque of an induction motor is the torque produced at the rotor. Hence $T = T_r$ where T_r is the rotor torque.

In case of D.C. motor torque = Armature Torque = T_a

$T_a = 0.159\phi Z I_a \left(\dfrac{P}{A}\right) \text{N·m}$

There fore $T_a = K\phi I_a$

[Where 0.159, Z, P and A are all constants)

Where ϕ is the flux produced by the filed winding which is pulsating in nature.

Similarly in case of an induction motor the torque is also proportional to the product of flux produced in stator and rotor current.

However there is another factor which is to be taken is power factor. Because in this case both flux and current are alternating in nature.

Therefore $T_r \propto \phi\, I_2 \cos \phi_2$

Where I_2 – Rotor Current

ϕ - flux produced in the stator.

ϕ_2 – The phase angle between rotor emf and rotor current (E_2 and I_2)

As $\phi \propto E_2$

Therefore $T_r = T \propto E_2\, I_2 \cos \phi_2$

$$T = K\,E_2\,I_2 \cos \phi_2$$

Starting Torque :

The torque developed by the motor at the instant of starting is called starting torque.

Let E_2 = Rotor emf per phase at stand still

R_2 = Rotor resistance / phase

X_2 = Rotor reactance / phase at stand still

$Z_2 = \sqrt{R_2^2 + X_2^2}$ = Rotor impedance / phase at stand still

Then $I_2 = \dfrac{E_2}{Z_2} = \dfrac{E_2}{\sqrt{R_2^2 + X_2^2}}$, $\cos \phi_2 = \dfrac{R_2}{Z_2} = \dfrac{R_2}{\sqrt{R_2^2 + X_2^2}}$

Stand still or starting torque $T_{st} = K\,E_2\,I_2 \cos \phi_2$

$$\text{Or } T_{st} = K\,E_2 \cdot \frac{E_2}{\sqrt{R_2^2 + R_2}} \cdot \frac{R_2}{\sqrt{R_2^2 + R_2}} = \frac{K\,E_2^2\,R_2}{R_2^2 + R_2}$$

If supply voltage V remains constant, then the flux ϕ and hence E_2 remain constant.

$$\text{Therefore } T_{st} = K_1 \frac{R_2}{R_2^2 + X_2^2}$$

$$\Rightarrow \text{Tst} = K_1 \frac{R_2}{Z_2^2}$$

Starting Torque of a Squirrel – cage Induction Motor :

The resistance of a squirrel cage motor is fixed and small as compared to its reactance which is very large especially at the start because at stand still, the frequency of the rotor currents equal the supply frequency. Hence the starting current I_2 of the rotor, though very large in magnitude, laggs by a very large angle E_2, with the result that the starting torque per ampere is very poor. Hence, such motors are not useful where the motor has to start against heavy loads.

Starting Torque of a slip-ring motor :

The starting torque of such motor is increased by improving its power factor by adding external resistance in the rotor circuit from the star connected rheostat, the rheostat resistance being progressively cut out as the motor gathers speed. Addition of external resistance, however increases the rotor impedance and so reduces the rotor current. At first, the effect of improved power factor predominates the current–decrearing effect of impedance. Hence, starting torque is increased. But after a certain point, the effect of increased impedance predominates the effect of improved power factor and so the torque starts decreasing.

Condition for maximum starting Torque :

As starting torque $T_{st} = \dfrac{K_2 R_2}{R_2^2 + X_2^2}$

From mathematics we know that differentiation of a maximum quantity = 0

D (T_{st}) = 0, when T_{st} = Maximum starting Torque

Therefore $\dfrac{d(T_{st})}{dR_2} = 0$

$$\Rightarrow \frac{d}{dR_2}\left(\frac{K_2 R_2}{R_2^2 + X_2^2}\right) = 0$$

$$\Rightarrow K_2 \frac{d}{dR_2}\left(\frac{R_2}{R_2^2 + X_2^2}\right) = 0$$

$$\Rightarrow \frac{d}{dR_2}\left(\frac{R_2}{R_2^2 + X_2^2}\right) = 0$$

$$\Rightarrow \frac{\left(R_2^2 + X_2^2\right)\cdot\dfrac{d}{dR_2}\cdot R_2 \;-\; R_2\dfrac{d}{dR_2}\left(R_2^2 + X_2^2\right)}{\left(R_2^2 + X_2^2\right)^2} = 0$$

$$\Rightarrow R_2^2 + X_2^2 \cdot 1 - R_2\left(2R_2 + 0\right) = 0$$

$$\Rightarrow R_2^2 + X_2^2 - 2R_2^2 = 0$$

$$\Rightarrow X_2^2 = R_2^2$$

$$\Rightarrow R_2 = X_2$$

Hence the starting torque will be maximum when Rotor resistance = Rotor Reactance.

Rotor EMF and Rotor reactance under running condition :

Rotor EMF : Let E_2 = Stand still rotor EMF / phase

X_2 = Stand still rotor reactance / phase

When rotor starts rotating, the relative speed between rotor and rotating flux in the stator starts decreasing.

$$\text{Slip (s)} = \frac{N_s - N_r}{N_s}$$

The rotor induced emf is directly proportional to this relative speed

i.e. $E_r \propto (N_s - N_r) E_2$

$$\Rightarrow E_r = K (N_s - N_r) E_2$$

$$\Rightarrow E_r = \frac{N_s - N_r}{N_s} \cdot E_2$$

Therefore $E_r = SE_2$

Rotor Reactance :

The frequency of the rotor current

$f_r = sf$

Therefore $X_r = 2\pi s f_r L$

$$\Rightarrow X_r = 2\,\pi s f L$$

$$\Rightarrow X_r = S\,(2\pi f L)$$

Therefore $X_r = S\,X_2$

Torque under running conditions :

As we know that starting torque $T_{st} = KE_2 I_2 \cos\phi_2$

Therefore $T_{st} \propto E_2\, I_2 \cos\phi_2$

So the torque under running condition $T_r \propto E_r\, I_r \cos\phi_r$

Where E_r = Rotor EMF/Phase under running condition

I_r = Rotor Current/Phase under running condition

$A_s\, E_r \propto \phi$

Therefore $T_r \propto \phi\, I_r \cdot \cos\phi_r$

$I_r = \frac{E_r}{Z_r}$ But $Z_r = R_2 + j\,X_r = R_2 + j\,SX_2$

$\text{Cos}\,\phi_r = \frac{R_2}{\sqrt{R_2^2 + (SX_2)^2}}$ and $I_r = \frac{SE_2}{\sqrt{R_2^2 + (SX_2)^2}}$

Therefore running torque $T_r \;\alpha\; E_r\, I_r$ as ϕ_r

Therefore $T_r \;\alpha\; \phi \frac{SE_2}{\sqrt{R_2^2 + (S \cdot X_2)^2}} \cdot \frac{R_2}{\sqrt{R_2^2 + (S \cdot X_2)^2}}$

$\Rightarrow$ $T_r \;\alpha\; \phi \frac{SE_2 R_2}{R_2^2 + (S \cdot X_2)^2}$

$\Rightarrow$ As $E_2 \;\alpha\; \phi$

Otenu $T_r \;\alpha\; \frac{SE_2^2 R_2}{R_2^2 + (S \cdot X_2)^2}$

Therefore $T_r = \frac{K_1 SE_2^2 R_2}{R_2^2 + (S \cdot X_2)^2}$

Torque under stand still condition :

Nr = 0 at stand still condition

$S = \frac{N_S - 0}{N_S} = 1$

Therefore torque under stand still condition

$Tr = \frac{K_1\, E_2^2\, R_2}{R_2^2 + X_2^2}$

Condition for maximum Torque under running condition :

The torque of a rotor under running condition

$T_r = \frac{K_1 SE_2^2 R_2}{R_2^2 + (S \cdot X_2)^2}$

The conditions for maximum torque may be obtained by differentiating the above equation w.r.t slip (s) and then putting it equal to zero.

Let $Y = \dfrac{1}{T_r}$ (For to make the differentiation easy)

Therefore $Y = \dfrac{R_2^2 + (SX_2)^2}{K_1 S E_2^2 R_2}$

$$\Rightarrow Y = \frac{R_2}{K_1 S E_2^2} + \frac{SX_2^2}{K_1 E_2^2 R_2}$$

For maximum torque under running condition $\dfrac{dY}{dS} = 0$

$$\Rightarrow \frac{d}{dS}\left(\frac{R_2}{K_1 S E_2^2}\right) + \frac{d}{dS}\left(\frac{SX_2^2}{K_1 E_2^2 R_2}\right) = 0$$

$$\Rightarrow \frac{d}{dS}\left(\frac{R_2}{S E_2^2}\right) + \frac{d}{dS}\left(\frac{SX_2^2}{E_2^2 R_2}\right) = 0$$

$$\Rightarrow \frac{\frac{dR_2}{dS}\cdot SE_2^2 \;-\; R_2\frac{d}{dS}\left(SE_2^2\right)}{\left(S\,E_2^2\right)^2} + \frac{\frac{d}{dS}\left(SE_2^2\right)\cdot E_2^2 R_2 \;-\; \frac{d}{dS}\,E_2^2 R_2\cdot\left(SX_2^2\right)}{\left(E_2^2 R_2\right)^2} = 0$$

$$\Rightarrow \frac{0 - \cdot\; E_2^2 R_2}{S^2\, E_2^4} + \frac{X_2^2 E_2^2 R_2 - 0}{E_2^4\;R_2^2} = 0$$

$$\Rightarrow \frac{-\,R_2 E_2^2}{S^2\,E_2^4} + \frac{X_2^2}{E_2^2\,R_2} = 0$$

$$\Rightarrow \frac{R_2}{S^2\,E_2^2} = \frac{X_2^2}{E_2^2 R_2}$$

$$\Rightarrow \frac{R_2}{S^2} = \frac{X_2^2}{R_2}$$

$$\Rightarrow R_2^2 = S^2 X_2^2$$

Therefore $\boxed{R_2 = SX_2}$

Hence the torque under running condition will be maximum when $R_2 = SX_2$

As the torque under running condition

$$T_r = \frac{K_1 S E_2^2 R_2}{R_2^2 + (SX_2)^2}$$

Putting the value $R_2 = SX_2$

Therefore $T_r = T_r$ (max) = $\dfrac{K\,SE_2^2 \cdot SX_2}{(SX_2)^2 + (SX_2)^2}$

$\Rightarrow T_r$ (max) = $\dfrac{KS^2E_2^2X_2}{2S^2X_2^2} = \dfrac{K\,E_2^2}{2X_2}$

Hence $\boxed{T_r\,(\max) = \dfrac{K\,E_2^2}{2X_2}}$

Relation between full load Torque and Maximum Torque :

As Torque (T) = $\dfrac{K_1\,SE_2^2\,R_2}{R_2^2 + (SX_2)^2}$

E_2 is practically constant

Hence T = $\dfrac{K_2\,SR_2}{R_2^2 + (SX_2)^2}$

Therefore T α $\dfrac{SR_2}{R_2^2 + (SX_2)^2}$

Taking full load slip as S_f at full load torque T_f

Therefore T_f α $\dfrac{S_f R_2}{R_2^2 + (SX_2)^2}$ ………. (I)

As $T_{max} = \dfrac{K\,E_2^2}{2X_2}$

T_{max} α $\dfrac{1}{2X_2}$ ……….. (II)

$$\frac{(i)}{(ii)} = \frac{T_f}{T_{max}} = \frac{S_f\,R_2}{R_2^2 + (S_f X_2)^2} \times \frac{2X_2}{1}$$

$$\frac{T_f}{T_{max}} = \frac{2S_f R_2 X_2}{R_2^2 + (S_f X_2)^2}$$

Dividing X_2^2 on both side

$$\Rightarrow \frac{T_f}{T_{max}} = \frac{2S_f \frac{R_2}{X_2}}{\frac{R_2^2}{X_2^2} + S_f^2}$$

Taking $\frac{R_2}{X_2} = a$

$$\Rightarrow \frac{T_f}{T_{max}} = \frac{2aS_f}{a^2 + S_f^2}$$

In general $\frac{\text{operating Torque}}{\text{Maximum Torque}} = \frac{2as}{s^2 + a^2}$

s – operating slip

Relation between starting Torque and Maximum Torque :

As $T_{st} = K \frac{R_2}{R_2^2 + X_2^2}$

$$\Rightarrow T_{st} \;\alpha\; \frac{R_2}{R_2^2 + X_2^2} \qquad \ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots \text{(i)}$$

But $T_{max} \;\alpha\; \frac{1}{2X_2}$ …………………. (ii)

$$\frac{\text{(i)}}{\text{(ii)}} = \frac{T_{st}}{T_{max}} = \frac{R_2}{R_2^2 + X_2^2} \times \frac{2X_2}{1}$$

$$\Rightarrow \frac{T_{st}}{T_{max}} = \frac{2R_2X_2}{R_2^2 + X_2^2}$$

$$\Rightarrow \frac{T_{st}}{T_{max}} = \frac{\frac{2R_2X_2}{X_2^2}}{\frac{R_2^2}{X_2^2} + \frac{X_2^2}{X_2^2}}$$

$$\Rightarrow \frac{T_{st}}{T_{max}} = \frac{\frac{2R_2}{X_2}}{\left(\frac{R_2}{X_2}\right)^2 + 1}$$

$$\Rightarrow \boxed{\frac{T_{st}}{T_{max}} = \frac{2a}{a^2+1}}$$

Relation between Torque and slip :

As Torque (T) = $\dfrac{KSE_2^2R_2}{R_2^2+(SX_2)^2}$

Taking Torque in Y axis and slip in X axis

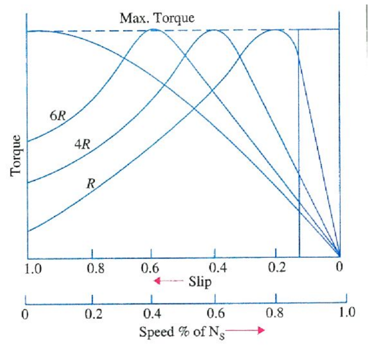


Fig. 1.5

At origin i.e. S = 0, torque T = 0

Therefore the curve starts from origin. At normal speed, closed to synoronism that is when N_r is very near to N_s, then slip is very nearly equal to zero.

Therefore $SX_2 << R_2$

$$\Rightarrow T \;\alpha\; \frac{SE_2^2R_2}{R_2^2} \qquad \{\text{Neglecting } (SX_2)^2\}$$

(Taking supply voltage constant so E_2 is also constant)

$$\Rightarrow T \;\alpha\; \frac{S}{R_2}$$

For a particular induction motor R_2 is constant.

Hence T α S

Therefore low valve of slip, torque is directly proportional to slip. Hence the curve is straight line for low valve of slip.

As slip increases the torque also increases and becomes maximum when = $R_2 = SX_2$

i.e. $S = \dfrac{R_2}{X_2}$

As the slip further increases (SX_2) becomes higher compare to R_2.

Hence R_2 can be neglected in compare to (SX_2)

$$\Rightarrow T \alpha \frac{S}{(SX_2)^2}$$

$$\Rightarrow T \alpha \frac{1}{SX_2^2}$$

Taking X_2 is constant for a particular induction motor

Therefore $T \; \alpha \; \dfrac{1}{S}$

So beyond the point of maximum torque any further increase in slip, results in decrease of torque.

Method of standing of Induction Motor

The operation of the squirrel cage induction motor is similar to transformer having short circuited on the secondary side.

Due to short circuited on the rotor circuit it will take heavy current when it is directly switched on. Generally when direct switched, take five to seven times of their full load current. This initial excessive current is objectionable, because it will produce large line voltage drop.

Hence it is not advisable to start directly motors of rating above 5 KW. But the starting torque of an induction motor can be improved by increasing the resistance of the rotor circuit. This is easily feasible in the case of slip ring induction motor but not in the case of squirrel cage motors. However, in their case, the initial in rush of current is controlled by applying a reduced voltage to the stator during the starting period, full normal voltage being applied when the motor has run up to speed.

Method of Starting of Squirrel Cage Motor :

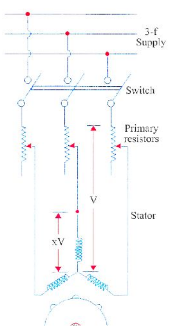
(1) Resistors Method

(2) Star – Delta Method

(3) Auto transformer Method

In the above methods, the supply voltage to the squirrel cage motor is reduced during starting.

1) Resistor Method :



In this method the resistors are connected in series with the stator phases, to give reduced voltage to the stator winding.

When resistors are connected in series with the stator phases, the current in the stator phases will reduce. If the voltage applied across the motor terminals is reduced by 50%, starting current is reduced by 50%.

Fig 1.6

When the motor starts running the resistances in the circuit is gradually cut out and full voltage is applied to the stator circuit. This method is useful for the smooth starting of small machines only.

2) Star – Delta Starter :

This method is used in the case of motors which are built to run normally with a delta connected stator winding. It consists of a two way switch which connects the motor in star for starting and then in delta for normal running.

At starting, when star connected, the applied voltage over each motor phases is reduced by afactor $1/\sqrt{3}$. Hence during starting, when motor is star connected it takes $1/\sqrt{3}$ times as much as starting current.

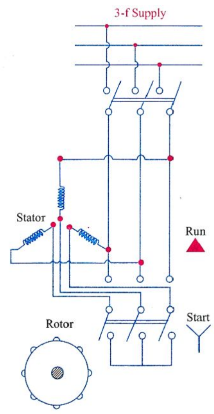


Fig 1.7

When the motor catches the speed 80% of its normal speed switch is changed to delta positions at that time $V_L = V_{ph}$.

Auto Transformer Method :

This starter is popularly known as auto starter in auto transformer the secondary side gets less voltage in compare to primary side.

As shown in the figure, at starting condition, a reduced voltage is applied across the mo terminals. When the motor catches the speed 80% of its normal speed, connections are changed to running position, then full supply voltage is applied across the motor.

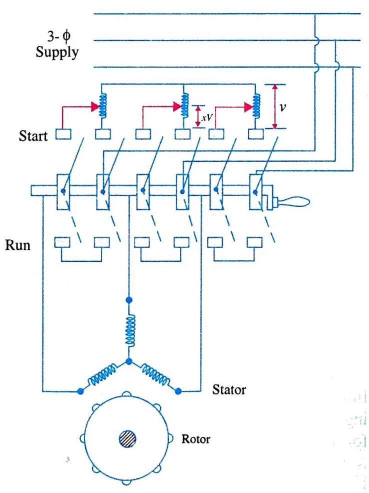


Fig 1.8

Most of the auto starters are provided with 3 – sets of taps so as to reduced the voltage to 80, 65 or 50 percent of line voltage.

Slip ring Motor :

Rotor Rheostat Method :

These motors are practically always started with full line voltage applied across the stator terminals. The value of starting current is adjusted by introducing a variable resistance in the rotor circuit.

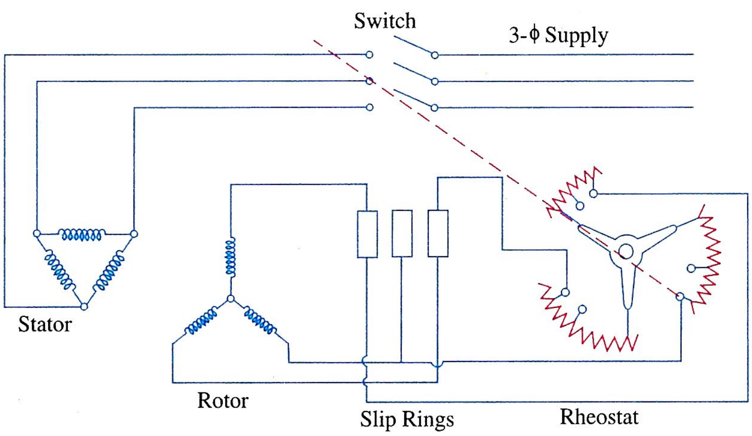


Fig 1.9

The controlling resistance is in the form of a rheostat, connected in star, the resistance being gradually cut – out of the rotor circuit, as the motor gathers speed

Speed Control of Induction Motor :

The speed of an induction motor can be changed under two main headings.

(i) Control from stator side

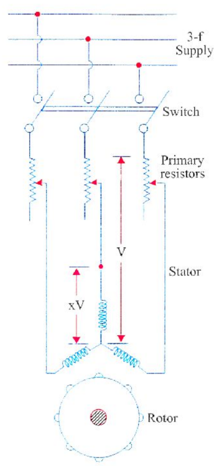
(ii) Control from Rotor side

(i) **Control from stator side :**

(a) By changing the applied voltage

(b) By changing the applied frequency

(c) By changing the no of stator poles.



(ii) **Control from Rotor side :**

(a) Rotor Rheostatic Control

(b) Cascade operation

Fig 1.10

(c) By injecting emf in the rotor circuit

By changing applied voltage :

This method is the easiest way for controlling speed of an induction motor. But this method is rarely used for the following reasons.

(i) A large change in voltage is required for a small change in speed.

(ii) Due to the connection of resistances in the stator phases, large power loss occurs at the resistors.

When the resistances are added in the stator circuit, voltage across the stator phase decreases.

As torque (T) = $\frac{KV^2R_2}{R_2^2 + X_2^2}$

$\Rightarrow$ Torque T = $K_1 V^2$

$\Rightarrow T \alpha V^2$

The torque depends on the supply voltage on the stator terminals, when V will decrease T will decrease hence speed will decrease.

By Charging the number of stator poles :

This method is easily applicable to squirrel cage motors because the squirrel cage rotor adopts it self to any reasonable number of stator poles.

The change in number of stator poles is achieved by having two more entirely independent stator windings in the same slots. Each winding gives a different number of poles and hence different synchronous speed.

Rotor Rheostatic Control :

This method is applicable to slip ring motors alone. The motor speed is reduced by introducing an external resistance in the rotor circuit.

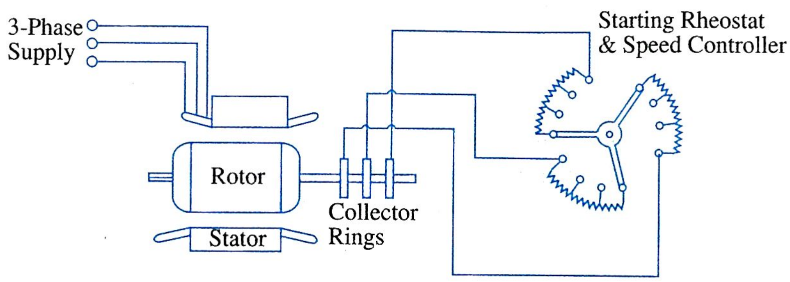


Fig 1.11

For this purpose the rotor starter may be used.

As torque (T) $\alpha \ \frac{S}{R_2}$

By increasing the motor resistance torque will decrease. Hence speed will decrease.

Motor Enclosures :

Enclosed and semi-enclosed motors are practically identical with open motors in mechanical construction and in their operating characterstics. Many different types of frames or enclosures are available to suit particular requirements. Some of the common type of enclosures are given below.

(i) Totally enclosed, Non ventilated type.

(ii) Splash – Proof type

(iii) Totally enclosed, Fan cooled type.

(iv) Cowl covered motor

(v) Protected Type

(vi) Drip – Proof Motors

(vii) Self (Pipe) Ventilated Type

(viii) Separately (Forced) Ventilated Type.

Induction Generator :

When the rotor of an induction motor runs faster than its synchronous speed at that time the induction motor runs as a generator called Induction generator. It converts the mechanical energy it receives into electrical energy is released by the stator.

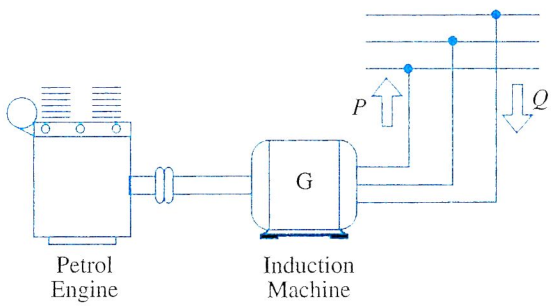


Fig 12

Figure shows a ordinary squirrel cage induction motor which is driven by a petrol engine and is connected to a 3 – phase line. As soon as motor speed exceeds its synchronous speed, it starts delivering active power P to the 3 – phase line. However, for creating its own magnetic field, it absorbs reactive power Q from the line to which it connected.

-0-

CHAPTER-II

(Alternators)

INTRODUCTION

An alternating voltage is generated in a single conductor/coil **rotating** in a uniform magnetic field with **stationary** field poles .Similarly, an alternating voltage will also be generated in a **stationary** conductor/coil when the field poles **rotate** past the conductor/coil , as it is the **relative motion** between the field and the conductor/coil that matters so far as emf induction in a conductor/coil is concerned. The wave shapes of voltage in both the cases are sinusoidal as the wave shape of magnetic flux is sinusoidal.

In D.C. generators , the field poles are **stationary** and the armature conductors **rotate**.The voltage generated in the armature conductors is of alternating nature. This generated alternating voltage is converted to a direct voltage at the brushes with the help of the commutator.

A.C. generators are usually called **Alternators.** They are also called Synchronous generators. Rotating machines that rotate at a speed fixed by the supply frequency and the number of poles are called synchronous machines.

A synchronous generator is a machine for converting mechanical power from a prime mover to ac electric power at a specific voltage and frequency .A synchronous machine rotates at a constant speed called the **synchronous speed**.Synchronous machines are usually of 3-phase type because of various advantages of 3-phase Generation,Transmission and Distribution.Large synchronous generators of several MVA ratings are used to generate bulk power at thermal ,hydro and nuclear power stations.

ADVANTAGES OF ROTATING FIELD ALTERNATOR

Most alternators have the rotating field and the stationary armature.The rotating-field type alternator has several advantages over the rotationg-armature type alternator.

(1) A stationary armature is more easily insulated for the high yoltage for which the alternator is designed. This generated voltage may be as high as 33KV.
(2) The armature windings can be braced better mechanically against high electro-magnetic forces due to large short-circuit currents when the armature windings are in the stator.
(3) The armature windings, being stationary, are not subjected to vibration and centrifugal forces.

(4) The output current can be taken directly from fixed terminals on the stationary armature without using slip rings, brushes, etc.

(5) The rotating field is supplied with direct current. Usually the field voltage is between 100 to 500 volts. Only two slip rings are required to provide direct current for the rotating field while at least three slip rings would be required for a rotating armature. The insulation of the two relatively low voltage slip rings form the shaft can be provided easily.

(6) The bulk and weight of the armature windings are substantially greater than the windings of the field poles. The size of the machine is, therefore, reduced.

(7) Rotating field is comparatively light and can be constructed for high speed rotation. The armatures of large alternators are forced cooled with circulating gas or liquids.

(8) The stationary armature may be cooled more easily because the armature can be made large to provide a number of cooling ducts.

SPEED AND FREQUENCY

The frequency of the generated voltage depends upon the number of field poles and on the speed at which the field poles are rotated. One compete cycle of voltage is generated in an armature coil when a pair of field poles (one north and one south pole) passes over the coil.

Let P =total number of field poles

P= pair of field poles

N= speed of the field poles in r.p.m.

n=speed of the field poles in r.p.s.

f= frequency of the generated voltage in Hz

Obviously $\frac{N}{60} = n$ ……………………………………………….. 1.1

and $\frac{P}{2} = p$ ………………………………………………………………1.2

In one revolution of the rotor, an armature coil is cut by $\frac{P}{2}$ north poles and $\frac{P}{2}$ south poles. Since one cycle is generated in an armature coil when a pair of field poles passes over the coil, the number of cycles generated in one revolution of the rotor will be equal to the number of pairs of poles. That is, Number of cycles per revolution = p

Also, number of revolutions per second=n

Now frequency=number of cycles per second

$$=\frac{number\ of\ cycles}{revolutons} X \frac{revolutons}{seconds}$$

f= p x n ………………………………………………………… 1.3

Since n =N/60 and p=p/2

$f=\frac{PN}{120}$ ……………………………………………………….. 1.4

Equation(1.2) and (1.4) give the relationship between the number of poles, speed and frequency.

SYNCHRONOUS SPEED

From Eq.(1.4)

$Ns=\frac{120\ f}{p}$ ………………………..(1.5)

Equation (1.5) shows that the rotor speed N bears a constant relationship with the field poles and the frequency of the generated voltage in the armature winding. The speed given by Eq. (1.5) us called synchronous speed N_s . A machine which runs at synchronous speed is called synchronous machine. Thus, a synchronous machine is an a.c. machine in which the rotor moves at a speed which bears a constant relationship to the frequency of the generated voltage in the armature winding and the number of poles of the machine .Table 1.1 gives the number of poles and synchronous speeds for a power frequency of 50H_z.

Table 1.1

Number of poles	Synchronous speed N_s in r.p.m.
2	3000
4	1500
6	1000
8	750
10	600
12	500

EXMPLE 1.1

Calculate the highest speed at which (a) 50H_Z (b) $60\ H_Z$ alternator can be operated.

Solution

Since it is not possible to have fewer than 2 poles, the minimum value of p= 2.

$$f = \frac{PNs}{120}$$

$$N_s = \frac{120f}{p}$$

For a minimum value of P the speed N will be a maximum.

(a) f= 50 Hz, p=2

$N_S = \frac{120X50}{2}$ = 3000r.p.m.(Ans.)

(b) f=60 Hz, p=2

$N_S = \frac{120X60}{2}$ = 3600r.p.m.(Ans.)

CONSTRUCTION OF THREE-PHASE SYNCHRONOUS MACHINES

Similar to other rotating machines, an alternator consists of two main parts namely, the stator and the rotor. The stator is the stationary part of the machine. It carries the armature winding in which the voltage is generated. The output of the machine is taken from the stator. The rotor is the rotating part of the machine. The rotor produces the main field flux.

STATOR CONSTRUCTION

The various parts of the stator include the frame, stator core, stator windings and cooling arrangement. The frame may be of cast iron for small-size machines and of welded steel type for large size machines. In order to reduce hysteresis and Eddy-current losses, the stator core is assembled with high grade silicon content steel laminations. A 3-phase winding is put in the short cut on the inner periphery of the stator as shown in Fig.1. The winding is star connected. The winding of each phase is distributed over slots . When current flows in a distributed winding it produces an essentially sinusoidal space distribution of emf.

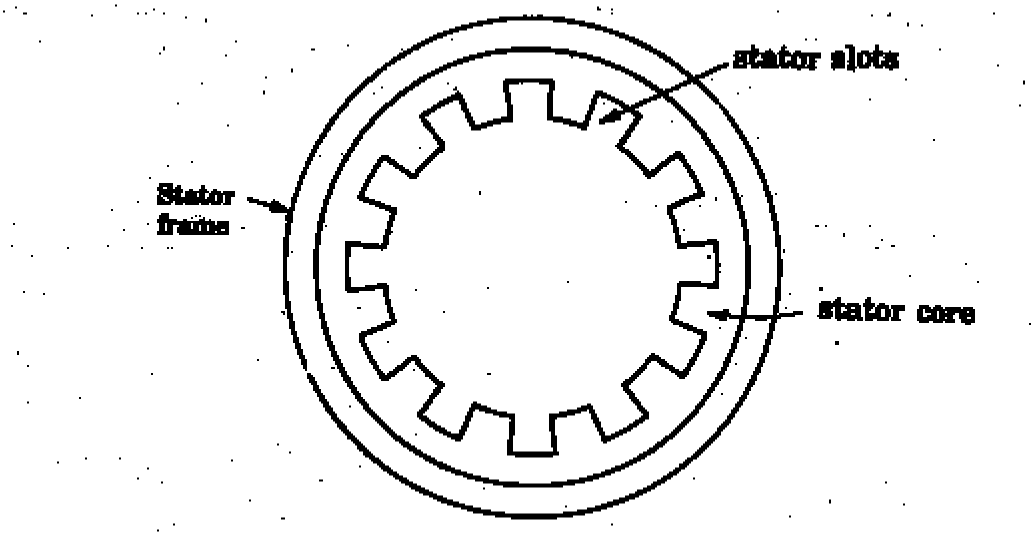


Figure 1.1 Alternator stator

ROTOR CONSTRUCTION

There are two types of rotor constructions namely, the salient-pole type and the cylindrical rotor type.

Salient-Pole Rotor

The term salient means '**protruding** 'or '**projecting**'. Thus, a salient-pole rotor consists of poles projecting out from the surface of the rotor cor. Figure 1.2 shows the end view of a typical 6-pole salient-pole rotor. Salient-pole rotors are normally used for rotors with four or more poles.

Since the rotor is subjected to changing magnetic fields, it is made of this steel laminations to reduce eddy current losses. Poles of identical dimensions are assembled by stacking laminations to the required length and then riveted together. After placing the field coil around each pole body, these poles are fitted by a dove-tail joint to a steel spider keyed to the shaft. Salient-pole rotors have concentrated winding on the poles. Damper bars are usually inserted in the pole faces to damp out the rotor oscillations during sudden change in load conditions. A salient-pole synchronous machine has a non-uniform air gap. The air gap is minimum under the pole centres and it is maximum in between the poles. The pole faces are so shaped that the radial air gap length increases from the pole centre to the pole tips so that the flux distribution in the air gap is sinusoidal. This will help the machine to generate sinusoidal emf.

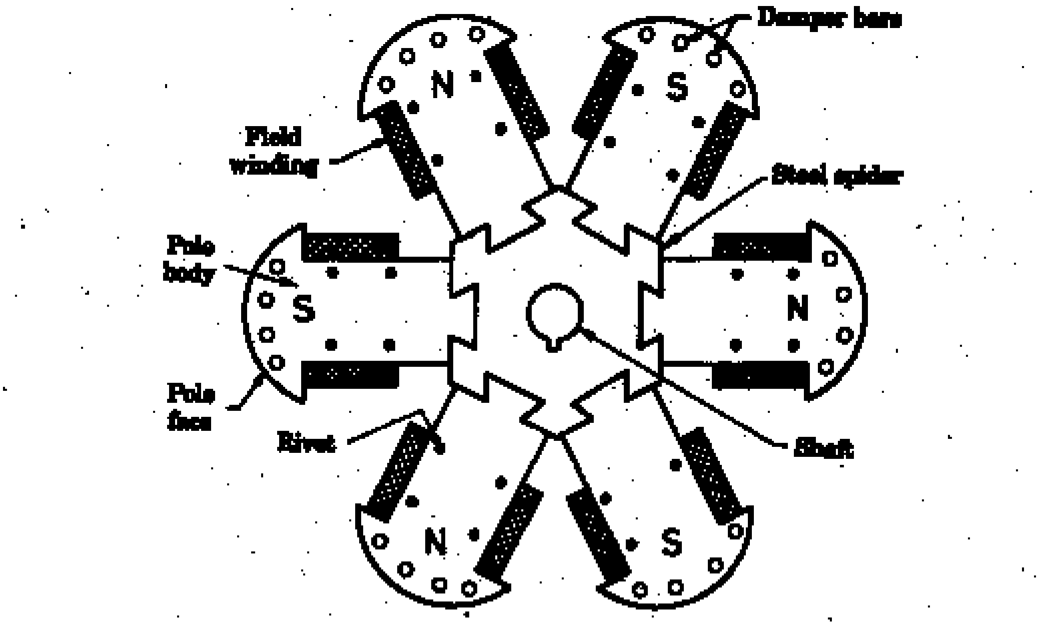


Figure 1.2 Six-pole salient-pole rotor

The individual field-pole windings are connected in series to give alternate north and south polarities. The ends of the field windings are connected to a dc source(a dc generator or a rectifier)through the brushes on the slip rings. The slip rings are metal rings mounted on the shaft and insulated from it. They are used to carry current to or from the rotating part of the machine (usually ac machine) via carbon brushes.

Salient-pole generators have a large number of poles at lower speeds. A salient-pole generator has comparatively a large diameter and a short axial length. The large diameter accommodates a large number of poles.

Salient-pole alternators driven by water turbines are called **hydro-alternators** or **hydro-generators**. Hydro-generators with relatively higher speeds are used with impulse turbines and horizontal configuration. Hydro-generators with lower speeds are used with reaction and Kaplan turbines and have vertical configuration.

Cylindrical Rotor

A cylindrical-rotor machine is also called a **non-salient pole rotor machine**. It has rotor so constructed that it forms a smooth cylinder. The construction is such that there are no physical poles to be seen as in the salient-pole construction. Cylindrical rotors are made from solid forgings of high grade nickel-chrome-molybdenum steel. In about two-third of the rotor periphery, slots are cut at regular intervals and parallel to the shaft. The dc field windings are accommodated in these slots. The winding is of distributed type. The un slotted portion of the rotor forms two (or four) pole faces. A cylindrical rotor machine has a comparatively small diameter and long axial length. Such a construction limits the centrifugal forces. Thus, cylindrical rotors are particularly useful in high-speed machines. The cylindrical rotor type alternator has two or four poles on the rotor. Such a construction provides a greater mechanical strength and permits more accurate dynamic balancing. The smooth rotor of the machine makes less windage losses and the operation is less noisy because of **uniform air gap**.

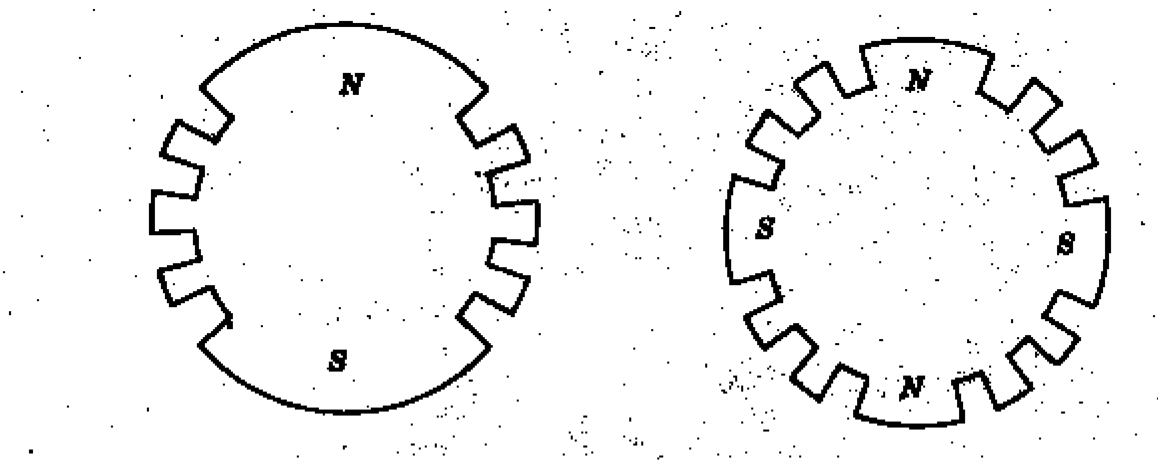


Figure 1.3 End views of two-pole and four pole cylindrical rotors

Figure 1.3 shows end views of 2-pole and 4-pole cylindrical rotors. Cylindrical rotor machines are driven by seam are gas turbines. Cylindrical rotor synchronous generators are called turbo-alternators or turbo-generators. Such machines have always horizontal configuration installation. The machines are built in a number of ratings from 10 MVA installed in super thermal power plants.

EXCITATION SYSTEMS FOR SYNCHRONOUS MACHINES

Excitation means production of flux by passing current in the field winding.

Direct current is required to excite the field winding on the rotor of the synchronous machines. For small machines, dc is supplied to the rotor field by a dc generator called exciter. This exciter may be supplied current by a smaller dc generator called **pilot exciter**. The main and pilot exciters are mounted on the main shaft of the synchronous machine(generator or motor).The dc output of the main exciter is giber to the field winding of the synchronous machine through brushes and slip rings. In smaller machines, the pilot exciter may be omitted, but this arrangement is not very sensitive or quick acting when changes of the field current are required by the synchronous machine.

For medium size machines a .c. exciters are used in place of d. c .exciters. A.C. exciters are three-phase a. c. generators. The output of a.c. exciter is rectified and supplied through brushes and slip-rings to the rotor winding of the main synchronous machine.

For large synchronous generators with ratings of few hundred megawatts, the excitation requirements become very large. The problem of conveying such amounts of power through high-speed sliding contacts becomes formidable. At present , large synchronous generators and synchronous motors are using brushless excitation systems. A brushless exciter is a small direct-coupled a . c. generator with its field circuit on the stator and the armature circuit on the rotor. The three-phase output of the ac exciter generator is rectified by solid-state rectifiers. The rectified output is connected directly to the field winding, thus eliminating the use of brushes and slip rings.

A brushless excitation system requires less maintenance due to absence of brushes and slip rings. The power loss is also reduced.

The d. c. required for the field of the exciter itself is sometimes provided by a small pilot exciter. A pilot exciter is a small a. c. generator with permanent magnets mounted on the rotor shaft and a three-phase winding on the stator. The permanent magnets of the pilot exciter produce the field current of the exciter. The exciter supplies the field current of the main machine. The use of a pilot exciter makes the excitation of the main generator completely independent of external supplies.

VOLTAGE GENERATION

The rotor of the alternator is run at its proper speed by its *prime mover*. The prime mover is a machine which supplies the mechanical energy input to the alternator. The prime movers used for a low and medium speed alternators are water wheels or hydraulic turbines. Steam and gas turbines are used as prime movers in large alternators and run at high speeds. The steam-turbine driven alternators are called *turboalternators* or *turbogenerators* . As the poles of the rotor move under the armature conductors on the stator, the field flux cuts armature conductors. Therefore voltage is generated in these conductors. This voltage is of alternating nature, since poles of alternate poles of alternate polarity successively pass by a given stator conductor. A 3-phase alternator has a stator with three sets of windings arranged so that there is a mutual phase displacement of 120^0. These windings are connected in star to provide a 3-phase output.

E. M.F. EQUATION OF AN ALTERNATOR

Φ = useful flux per pole in webers (Wb)

P= total number of poles

Z_p = total number of conductors or coil sides in sries per phase

T_p= total number of coils or turns per phase

n= speed of rotation of rotor in revolutions per second(r. p. s)

f= frequency of generated voltage(H_Z)

Since the flux per pole is Φ ,each stator conductor cuts a flux PΦ.

The average value of generated voltage per conductor

$$=\frac{flux\ cut\ per\ revolution\ in\ Wb}{time\ taken\ for\ one\ revolution\ in\ seconds}$$

Since n revolutions are made in one second, one revolution will be made in 1/n second.

Therefore the time for one revolution of the armature is 1/n second. The average voltage generated per conductor

$$E_{av}/\ \text{conductor} = \frac{P\Phi}{1/n} = np\Phi\ \text{volts} \quad ……………………………..1.6$$

We know that $f = \frac{PN}{120} = \frac{Pn}{2}$ …………………………………….1.7

$$Pn = 2f$$

Substituting the value of Pn in Eq.(1.6), we get

$$E_{av}/\ \text{conductor} = 2f\Phi \quad …………………………………………..1.8$$

Since there are Z_p conductors in series per phase, the average voltage generated per phase is given by

$$E_{av}/\ \text{phase} = 2f\Phi Z_p \quad ………………………………………………………………….. 1.9$$

Since one turn or coil has two sides, Z_p =2T_p, and the expression for the average generated voltage per phase can be written as

$$E_{av}/\ \text{phase} = 4f\Phi T_p \quad ……………………………………………………………….1.10$$

For the voltage wave, the form factor is given by

$$k_f = \frac{r.m.s.value}{average\ value}$$

For a sinusoidal voltage, k_f= 1.11. Therefore , the r.m.s. value of the generated voltage per phase can be written as

$E_{r.m.s}$/phase =k_fxE_{av}/phase= 1.11x4fΦT_p

=4.44 fΦT_p

The suffix r. m. s. is usually deleted , The r.m.s.value of the generated voltage per phase is given by

E_p = 4.44fΦT_p ………………………………………………………………. 1.11

Equation(1.11) has been derived with the following assumptions :

(a) Coils have got full pitch.

(b) All the conductors are concentrated in one stator slot.

ARMATURE WINDINGS

The winding through which a current is passed to produce the main flux is called the *field winding*. The winding in which voltage is induced is called *armature winding*. For synchronous machines the field windings are on the rotor.Therefore,the terms **rotor windings** and **field windings** are used interchangeably. Also,the armature windings are on the stator. Therefore, the term stator windings and armature windings are used interchangeably.

Some basic terms related to the armature winding are defined as follows :

A TURN consists of two conductors connected to one end by an end connector.

A coil is formed by connecting several turns in series .

The turn ,Coil and windings are shown schematically in figure 1.4.

The beginning of the turn, or coil, is identified by the symbol S(Start) and the end of the turn or coil by the symbol F(Finish).

The concept of electrical degrees is very useful in the study of machine. If

θ_{mech} = mechanical degrees or angular measure in space

θ_{elect} = electrical degrees or angular measure in cycles

For a P-pole machine electrical degree is defined as follows:

$$\theta_{elect} = \frac{p}{2}\ \theta_{mech}$$

The advantage of this notation is that expressions written in terms of electrical angles apply to machines having any number of poles.

The angular distance between the centres of two adjacent poles on a machine is known as pole pitch or pole span.

One pole pitch $=180^0_{ed} = \frac{360^0_{md}}{P}$

Regardless of the number of poles in the machine , a pole pitch is always 180 electrical degrees.

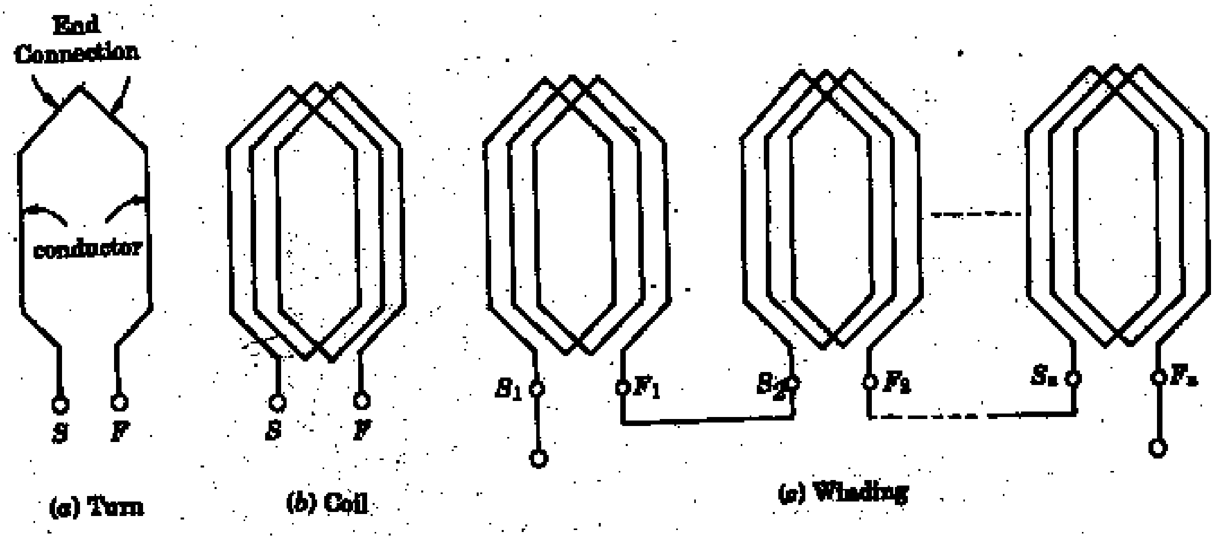


Figure 1.4 Turn , coil and winding

COIL-SPAN FACTOR OR PITCH FACTOR

The distance between the two sides of a coil is called the *coil span* or coil pitch.The angular distance between the central line of one pole to the central line of the next pole is called *pole pitch*.a pole pitch is always 180 electrical degrees regardless of the number of poles on the machine.A coil having a span equal to 180 degree electrical is called a *full-pitch* coil as shown in figure 1.5(a).

A coil having a span less than 180 electrical degrees is called a short-pitch coil or fractional pitch coil.It is also called a chorded coil.A stator winding using fractional-pitch coils is called a *chorded* winding.If the span of the coil is reduced by an angle of $\propto$ elctrical degrees,the coil span will be (180-$\propto$) electrical degrees as shown in figure 1.6(a).

In case of full-pitch coil two coil sides span a distance exactly equal to the pole pitch of 180 electrical degrees .As a result, the voltage generated in a full-pitch coil is such that the coil side voltages are in phase as shown in figure 1.5(b).Let $\mathbf{E_{C1}}$ & $\mathbf{E_{C2}}$ be the voltages generated in the coil sides and $\mathbf{E_C}$ the resultant coil voltage. Then

$$\mathbf{E_C} = \mathbf{E_{C1}} + \mathbf{E_{C2}}$$

Let **|Ec1|=|Ec2|=E1**

Since $\mathbf{E_{C1}}$ & $\mathbf{E_{C2}}$ are in phase,the resultant coil voltage $\mathbf{E_C}$ is equal to their arithmetic sum.

$$\mathbf{E_C} = \mathbf{E_{C1}} + \mathbf{E_{C2}} = 2\ \mathbf{E_1}$$

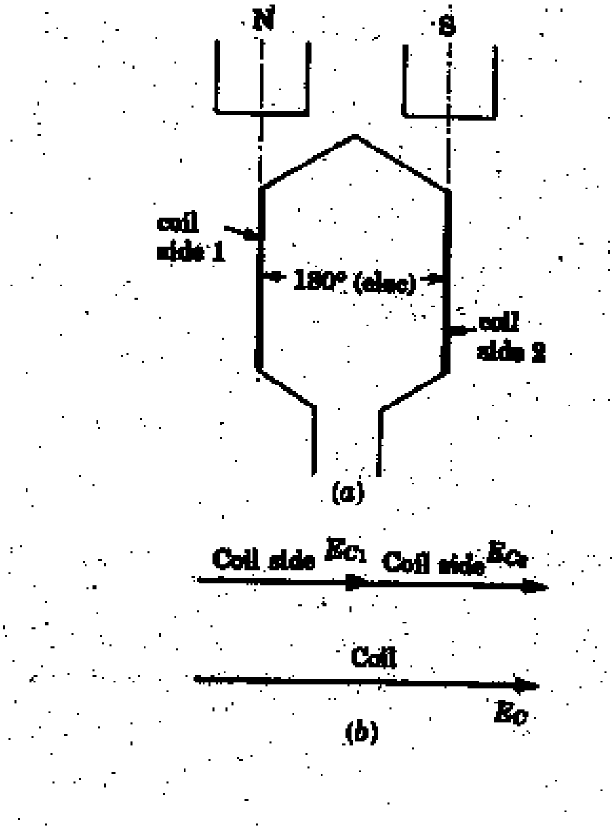


Fig. 1.5

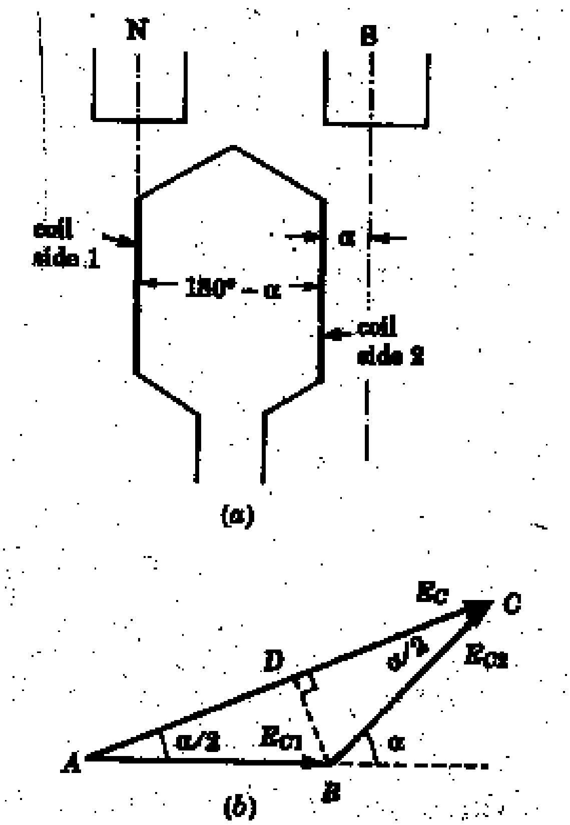


Fig. 1.6

If the coil span of a single coil is less than the pole pitch of 180^0 (elec.), the voltage generated in each coil side are *not* in phase.The resultant coil voltage E_C is equal to the *phasor sum* of E_{C1} & E_{C2}.

If the coil span is reduced by an angle $\propto$ elctrical degrees,the coil span is be (180-$\propto$) electrical degrees.The voltage generated E_{C1} & E_{C2} in two coil sides will be out of phase w.r.t each other by an angle $\propto$ *electrical degree as shown in figure* 1.6(b) . the phasor sum of E_{C1} & E_{C2} is E_C (=AC).

The coil span factor or pitch factor k_C is defined as the ratio of voltage generated in the short-pitch coil to the voltage generated in the full-pitch coil. The coil span factor is also called the *chording factor*.

$$k_C = \frac{Actual\ voltage\ generated\ in\ the\ coil}{voltage\ generated\ in\ the\ coil\ of\ span\ 180^0\ electrical}$$

$$\frac{Phasor\ sum\ of\ voltages\ of\ two\ coil\ sides}{arithmetic\ sum\ of\ the\ voltages\ of\ two\ coil\ sides}$$

$$= \frac{AC}{2AB} = \frac{2AD}{2AB} = \cos\frac{\propto}{2}$$

$$k_C = \cos\frac{\propto}{2} \quad \ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots \quad 1.12$$

For full-pitch coil, $\propto$=0, so, $\cos\frac{\propto}{2}$ =1 and k_C=1.For a short-pitch coil k_C< 1.

Advantages of short pitching or chording

1. Shortens the ends of the winding and therefore there is aa saving in the conductor material.
2. Reduces the effects of distorting harmonics, and thus the wave form of the generated voltage is improved and making it approach a sine wave.

DISTRIBUTION FACTOR OR BREADTH FACTOR k_d

In a concentrated winding ,the coil sides of a given phase are concentrated in a single slot under a given pole. The individual coil voltages induced are in phase with each other.These voltages may be added arithmetically. In order to determine the induced voltages induced per phase, a given coil voltage is multiplied by the number of series –connected coils per phase. In actual practice, in each phase, coils are not concentrated in a single slot, but are distributed in a number of slots in space to form a polar group under each pole. The voltages induced in coil-sides constituting a polar group are not in phase but differ by an angle equal to the angular displacement β of the slots.The total voltage induced in any phase will be the phasor sum of the individual coil voltages.

The distribution factor or breadth factor is defined as the ratio of the actual voltage obtained to the possible voltage if all the coils of a polar group were concentrated in a single slot.

$$k_d = \frac{\text{Phasor sum of coil voltages per phase}}{\text{arithmetic sum of coil voltages per phase}} \qquad \text{..................1.13}$$

Let m = the slots per pole per phase ,that is slots per phase belt

$$m = \frac{\text{slots}}{\text{Poles x Phases}} \qquad \text{......................................1.14}$$

β = Angular displacement between adjacent slots in electrical degrees

$$\beta = \frac{180^0}{\text{slots/pole}} = \frac{180^0 \times \text{poles}}{\text{slots}} \qquad \text{......................................1.15}$$

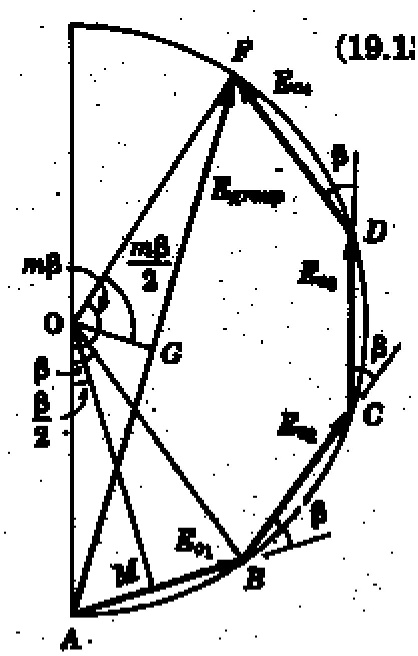


Fig.`1.7

Thus, one phase of the winding consists of coils arranged in m consecutive slots. Voltage E_{c1}, E_{c2} & E_{c3} ,…….. are the individual coil voltages .Each coil voltage E_c will be out of phase with the next coil voltage by the slot pitch β .Figure 1.7 shows the voltage polygon of the induced voltages in the four coils of a group(m=4).The voltages E_{c1}, E_{c2}, E_{c3} , and E_{c4} are represented by phasors AB,BC,CD and DF respectively in figure 1.7.Each of these phasors is a chord of a circle with center o and subtends an angle β at O.The phasor sum AF, representing the resultant winding voltage ,subtends an angle m β at the center.

Arithmatic sum of individual coil voltages

= m $\mathbf{E_c}$=m AB=m (2AM)

=2m sin Aom = 2m OA sin $\frac{\beta}{2}$

$$Phasor\ sum\ of\ individual\ coil\ voltages = AF = 2AG = 2OA \sin AOG = 2OA\ sin\frac{m\beta}{2}$$

$$k_d = \frac{Phasor\ sum\ of\ coil\ voltage\ per\ phase}{arithmetic\ sum\ of\ coil\ voltages\ per\ phase} = \frac{2\ OA \sin\frac{m\beta}{2}}{2\ OA\ m\ sin\frac{\beta}{2}}$$

$$k_d = \frac{\sin\frac{m\beta}{2}}{m\ sin\frac{\beta}{2}} \qquad \text{………………………………..1.16}$$

It is to be noted that the distribution factor k_d for a given number of phases is dependent only on the number of distributed slots under a given pole.It is independent of the type of the winding,Lap or wave,or the number of coil,etc. As the number of slots per pole increases the distribution factor decreases.

ACTUAL VOLTAGE GENERATED

Taking the oil span factor and the distribution factor into acount ,the actual generated voltage per phase is given by Ep =4.44 $K_c K_d f \Phi T_p$

………………………………………………………….. 1.17

Equation (1.17) is called the complete emf equation of an alternator.

The quantity $(K_c K_d T_p)$ is sometimes called effective turns per phase T_{ep} .

$$T_{ep} = K_c K_d T_p \qquad \text{…………………………………..1.18}$$

It ia smaller than the actual number of turns per phase due to fractional pitch coils and due to distribution of winding over several slots under each pole.

The coil span factor and distribution factor of a winding are sometimes combined into a single winding factor K_w which is the product of K_c and K_d. That is

$$K_w = K_c K_d \quad \ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots 1.19$$

For a star connected alternator , the line voltage is $\sqrt{3}$ times the phase voltage .

$$E_L = \sqrt{3}\, E_{ph}$$

Alternative terms for the voltage E are

Open circuit voltage per phase

No-Load voltage per phase

Excitation voltage per phase

Internal voltage per phase

Voltage behind synchronous reactance per phase

The angle between the terminal voltage V and the internal voltage E is the machine angle or rotor angle δ .

Example 1.2

A 3-phase , 50 Hz , 8-pole alternator has a star-connected winding with 120 slots and 8 conductors per slot.The flux per pole is 0.05 Wb, sinusoidally distributed.Determine the phase and line voltages.

Solution :

Let us take the full-pitch coil,

So, $\propto = 0^0$, $K_c = \cos\frac{\propto}{2} = \cos 0^0 = 1$

m = $\frac{slots}{Poles\ x\ Phase} = \frac{120}{8\times3}$ = 50 Hz

$$\beta = \frac{180^0 \times poles}{slots} = \frac{180^0 \times 8}{120} = 12$$

$$Kd = \frac{\sin\frac{m\beta}{2}}{m \sin\frac{\beta}{2}} = \frac{\sin\frac{5\times12}{2}}{5\ \sin\frac{12}{2}} = 0.9567$$

Total number of conductors = conductors per slot X number of slots = 8 x 120 = 960

Conductors per phase = ,$Z_p = \frac{960}{3} = 320$

Generated voltage per phase = $E_p = 2.22\ K_c K_d f \Phi Z_p$

$= 2.22 \times 1 \times 0.9567 \times 50 \times 0.05 \times 320 = 1699$ volts

Generated Line voltage = $E_L = \sqrt{3}\,E_{ph} = \sqrt{3} \times 1699 = 2942.8$ volts.

ARMATURE LEAKAGE REACTANCE

In a n ac machine ,any flux set up by the load current which does not contribute to the usuful flux of the machine is a leakage flux. The effect of this leakage flux is to set up a self-induced emf in the armature windings.

The leakage fluxes may be classified as follows :

1. **Slot leakage**
2. **Tooth head leakage**
3. **Coil-end or over-hang leakage**

The voltages induced in the armature windings by the air-gap flux is called the air-gap voltages.

The leakage fluxes also induce voltages in the armature windings .These are taken into account by introduction of leakage reactance drops. Most of the reluctances of the magnetic circuits for armature leakage fluxes are due to air paths. The fluxes are therefore nearly proportional to the armature currents producing them and are in phase with these currents. For this reason, the voltages they induce in the armature windings can be taken into account by the use of constant leakage reactances for the phases, which multiplied by the phase currents , give the component voltages induced in the phases of the leakage flux. These voltages are the leakage reactance drops and lead the currents producing them by 90^0.

ARMATURE REACTION

When load current flows through the armature windings of an alternator, the resulting mmf produces flux. This armature flux reacts with the main-pole flux, causing the resultant flux to become either less than or more than the original main flux. The effect of the armature(stator) flux on the flux produced by the rotor field poles is called **armature reaction**. The armature reaction flux is constant in magnitude and rotates at synchronous speed. The armature reaction depends upon the power factor of the load. If the armature reaction flux is assumed to act independently of the main field flux, it induces a voltage in each phase which lag the respective phase currents by 90^0. It is to be noted that armature reaction effects are seen to be an essential part of the torque producing mechanism.

Two things are worth noting about the armature reaction in an alternator. First,the armature flux and the flux produced by rotor ampere-turns rotate at the same speed (synchronous speed) in the same direction and,therefore,the two fluxex are fixed in space relative to each other.Secondly,the modification of flux in the air-gap due to armature flux depends on the magnitude of stator current and on the power factor of the load.It is the load power factor which determines whether the armature flux distorts,opposes or helps the flux produced by rotor ampere-turns.To illustrate this important point,the following three cases may be considered :

(i) When load p.f. is unity

(ii) When load p.f. is zero lagging

(iii) When load p.f. is zero leading

When load p.f. is unity : Figure 1.7.1(i) shows an elementary alternator on no-load.since the armature is on open-circuit, there is no stator current in the flux due to rotor current is distributed symmetrically in the air-gap as shown in fig.1.7.1(i).Since the direction of the rotor is assumed clockwise, the generated e.m.f in phase R_1R_2 is at its maximum and is towards the paper in the conductor R_1 and outwards in conductor R_2. No armature flux is produced since no current flows in the armature winding.

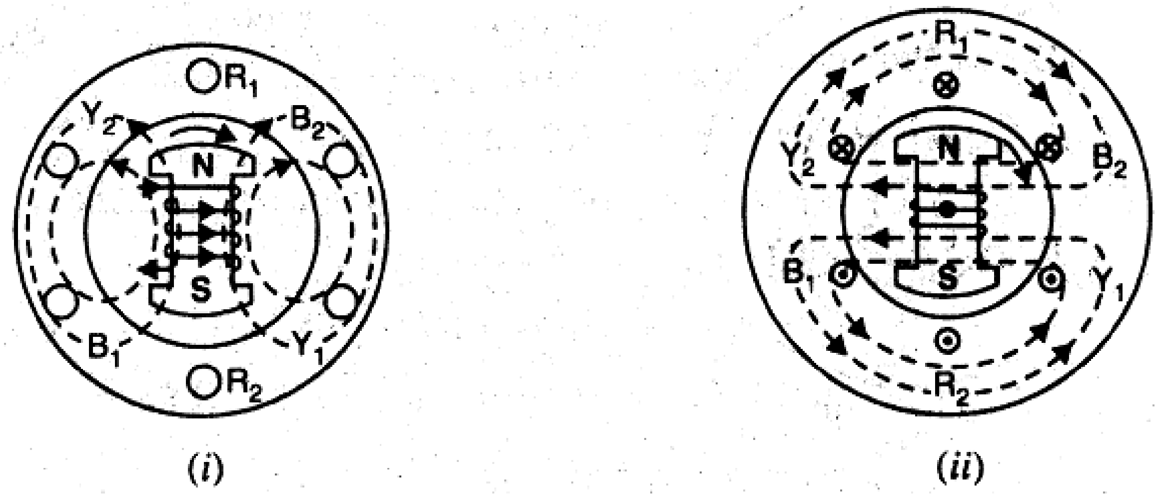


Fig.1.7.1

Fig.1.7.1(ii) shows the effect when a resistive load (unity p.f.) is connected across the terminals of the alternator. According to right-hand rule, the current is inwards in the conductors under N-pole and outwards in the conductors under S-pole. Therefore ,the armature flux is clockwise due to currents in the top conductors and anticlockwise due to current in the bottom conductors. Note that armature flux is at 90^0 to the main flux(due to rotor current) and is behind the main flux. In this case, the flux in the air-gap is distorted but not weakened. Therefore,at unity p.f. , the effect of armature reaction is merely to distort the main field;there is no weakening of the main field and the average flux practically remains the same. Since the magnetic flux due to stator currents(i.e. armature flux) rotates synchronously with the rotor, the flux distortion remains the same for all positions of the rotor.

When load p.f. is zero lagging :

when a pure inductive load(zero p.f. lagging) is connected across the terminals of the alternator ,current lags behind the voltage by 90^0.This means that current will be maximum at zero e..m.f. and vice-versa.

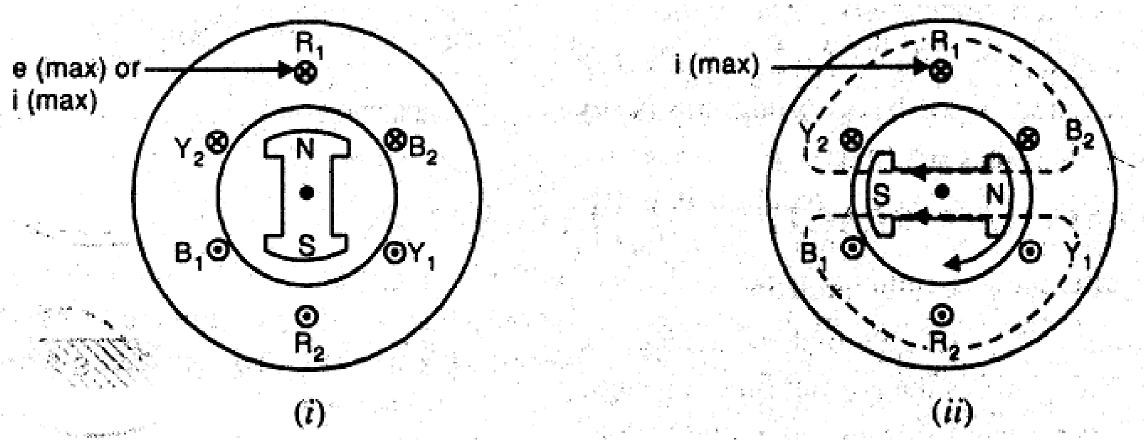


Fig.1.7.2

Fig.1.7.2(i) shows the condition when the alternator is supplying resistive load. Note that e.m.f. as well as current in phase R_1R_2 is maximum in the position shown.When the alternator is supplying a pure inductive load, the current in phase R_1R_2 will not reach its maximum value until N-pole advanced 90^0 electrical as shown in fig.1.7.2(ii). Now the armature flux is from right to left and field flux is from left to right. All the flux produced by armature current (i.e.,armature flux) opposes the field flux and , therefore, weakens it .In other words, armature reaction is directly de-magnetising. Hence, at zero p.f. lagging ,the armature reaction weakens the main flux. This causes a reduction in the generated e.m.f..

When load p.f. is zero leading :

when a pure capacitive load(zero p.f. leading) is connected across the terminals of the alternator, the current in armature windings will lead the induced e.m.f. by 90^0.Obviously, the effect of armature reaction will be the reverse that for pure inductive load. Thus, armature flux now aids the main flux and the generated e.m.f. is increased.

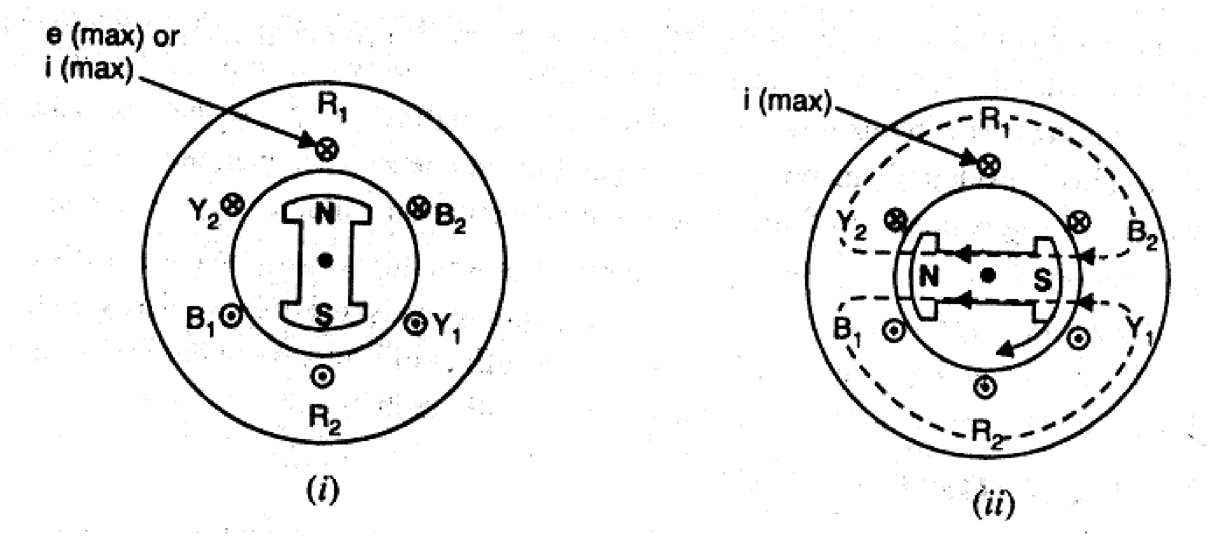


Fig.1.7.3

Fig.1.7.3(i) shows the condition when the alternator is supplying resistive load. Note that e.m.f. as well as current in phase R_1R_2 is maximum in the positon shown. When the alternator is supplying a pure capacitive load , maximum current in phase R_1R_2 will occur

90^0 *electrical* before the occurrence of maximum induced e.m.f.Therefore,maximum current in phase R_1R_2 will occur if the position of the rotor remains 90^0 behind as compared to its position under resistive load. This is illustrated in fig.1.7.3(ii).It is clear that armature flux is now in the same direction as the field flux and therefore, strengthens it. This causes an increase in the generated voltage. Hence at zero p.f. leading ,the armature reaction strengthens the main flux.

For intermediate values of p.f., the effect of armature reaction is partly distorting and partly weakening for inductive loads. For capacitive loads, the effect of armature reaction is partly distorting and partly strengthening. Note that in practice, loads are generally inductive.

Summary :

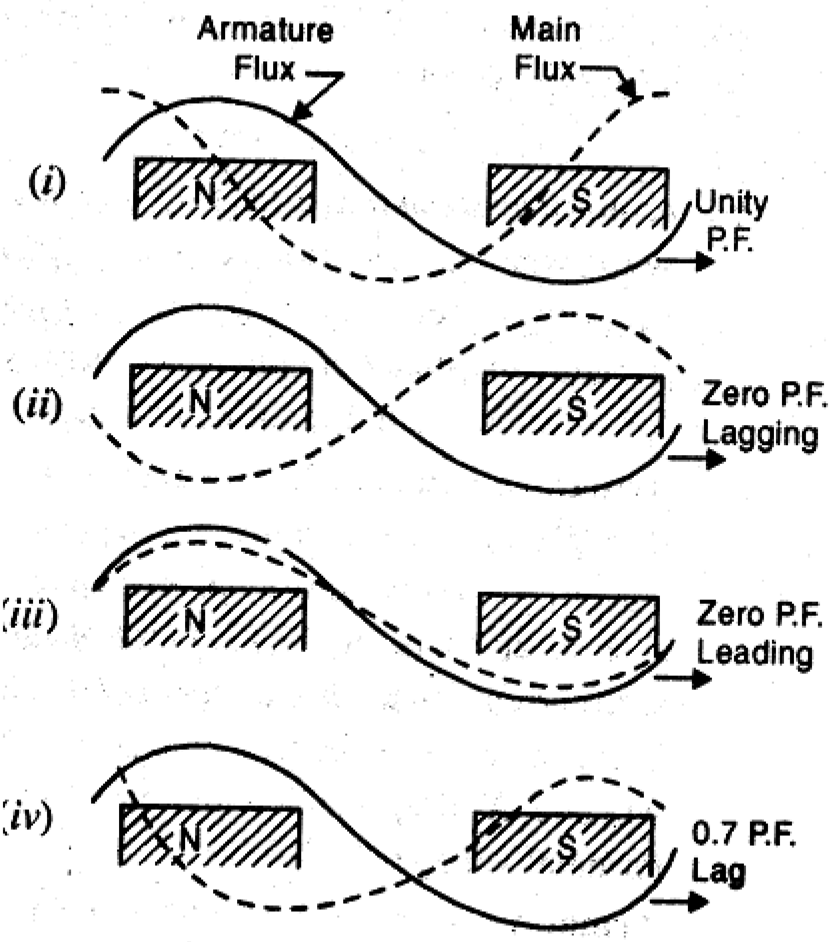


Fig.1.7.4

When the alternator is loaded, the armature flux modifies the air-gap flux .Its angle(electrical) with respect to main flux depends on the load p.f..This is illustrated in Fig.1.7.4.

(a) **When the load p.f. is unity** : the effect of armature reaction is wholly distorting. In other words, the flux in the air-gap is distorted but not weakened. As shown in Fig.1.7.4(i), the armature flux is 90^0 electrical behind the main flux. The result is that flux is strengthened at the trailing pole tips and weakened at the leading pole tips. However, the average flux in the air-gap practically remains unaltered.

(b) **When the load p.f. is zero lagging** : the effect of armature reaction is wholly demagnetising. In other words ,the flux in the air-gap is weakened. As shown in Fig.1.7.4(ii),the wave representing the main flux is moved backwards through 90^0 electrical so that it is in direct opposition to the armature flux. This considerably reduces the air-gap flux and hence the generated e.m.f..To keep the value of the generated e.m.f. the same,the field excitation will have to be increased to compensate for the weakening of the air-gap flux.

(c) **When the load p.f. is zero leading**: the effect of armature reaction is wholly magnetising. In other words ,the flux in the air-gap is increased. As shown in Fig.1.7.4(iii),the wave representing the main flux is now moved forward through 90^0 electrical so that it aids the armature flux. This considerably increases the air-gap flux and hence the generated e.m.f.. To keep the value of the generated e.m.f. the same,the field excitation will have to be reduced .

(d) **For intermediate values of load p.f.**: the effect of armature reaction is partly distorting and partly weakening for inductive loads. For capacitive loads, the effect is partly distorting and partly strengthening .Fig.1.7.4(iv) shows the effect of armature reaction for an inductive load. In practice, load on the alternator is generally inductive.

For the intermediate values of p.f. , the effect of armature reaction is partly distorting and partly weakening for inductive loads.For capacitive loads the effect of armature reaction is partly distorting and partly strengthening. Note that in practice, the loads are generally inductive.

SYNCHRONOUS IMPEDANCE

The actual generated voltage consists of the summation of two component voltages. One of these component voltages is the voltage that would be generated if there were no armature reaction. It is the voltage that would be generated because of only the field excitation. This component of the generated voltage is called the excitation voltage, E_{exc} .

The other component of the generated voltage is called the armature reaction voltage, E_{AR}. This is the voltage that must be added to the excitation voltage to take care of the effect of armature reaction upon the generated voltage

$$E_a = E_{exc} + E_{AR} \quad \ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots 1.20$$

Since armature reaction results, in a voltage effect in a circuit caused by change in flux by current in the same circuit ,its effect is of the nature of an inductive reactance .Therefore, E_{AR} is equivalent to a voltage of inductive reactance and

$$E_{AR} = -j\, X_{AR} I_a \quad \ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots 1.21$$

The inductive reactance X_{AR} is a ficticitous reactance which will result in a voltage in the armature circuit to account for the effect of armature reaction upon the voltage relations of the armature circuit. Therefore, armature reaction voltage can be modeled as an inductor in series with the internal generated voltage.

In addition to the effects of armature reaction, the stator winding also has a self-inductance and a resistance.

Let L_a = Self-inductance of stator winding

X_a = Self-inductive reactance of stator winding

R_a=Armature (stator) resistance

The terminal voltage V is given by

$$V = E_a - jI_a X_{AR} - jI_a X_a - I_a R_a$$

Where $I_a R_a$ = armature resistance drop

$I_a X_a$ = armature leakage reactance drop

$I_a X_{AR}$ = armature reaction voltage

The armature reaction effects and the leakage flux effects in the machine are both represented by inductive reactances . Therefore, it is customary to combine them in to a single reactance, called the synchronous reactance of the machine, X_s.

$$X_s = X_{AR} + X_a \quad \ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots 1.22$$

$$\therefore \qquad V = E_a - j\,X_s I_a - R_a I_a$$

$$\text{Or} \quad V = E_a - (R_a + jX_s)I_a \quad \ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots 1.23$$

$$V = E_a - Z_s I_a \quad \ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots 1.24$$

$$\text{Where} \quad Z_s = R_a + jX_s \quad \ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots 1.25$$

The impedance Z_s is called the *synchronous impedance*.

The synchronous reactance X_s is the ficticious reactance employed to account for the voltage effects in the armature circuit produced by the actual armature leakage reactance and by the change in air-gap flux caused by the armature reaction.

Similarly,the synchronous impedance Z_s is a ficticious impedance employed to account for the voltage effects in the armature circuit produced by the actual armature resistance,the actual armature leakage reactance and the change in air-gap flux caused by the armature reaction.

EQUIVALENT CIRCUIT AND PHASOR DIAGRAMS OF A SYNCHRONOUS GENERATOR

The equivalent reactance of a synchronous generator is shown in figure 1.8(a). It is redrawn in Fig.1.8(b) by taking

$$X_s = X_{AR} + X_a$$

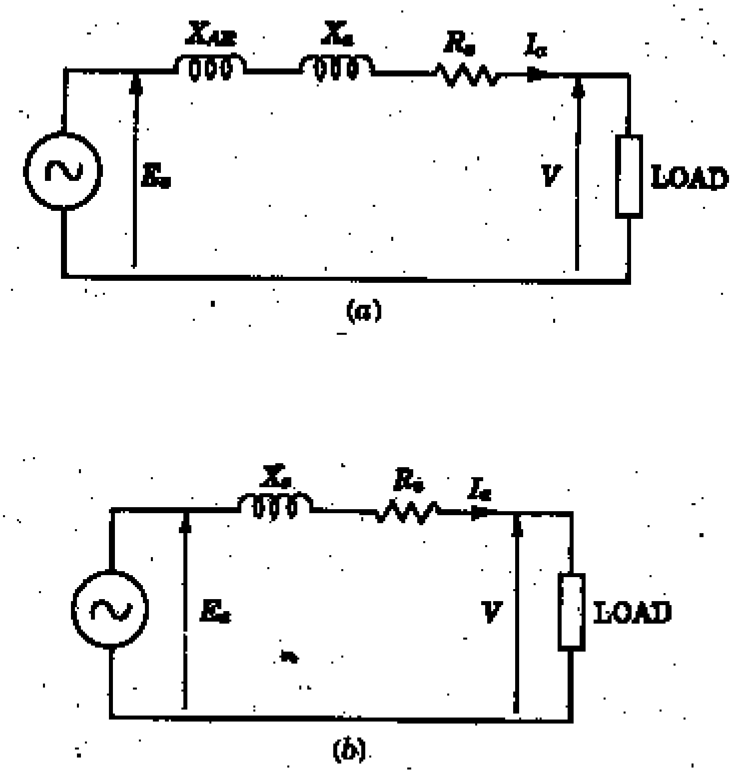


Fig.1.8 Equivalent circuit of a synchronous generator

a) Lagging power factor Cos Φ

Figure 1.9(a) shows the phasor diagram for lagging p.f. load. The power factor is Cos Φ lagging.In this diagram the terminal voltage V is taken as reference phasor along OA such that OA =V. For lagging power factor Cos Φ, the direction of armature current $\mathbf{I_a}$ lags behind the voltage **V** by an angle Φ along OB ,where OB = $\mathbf{I_a}$ .The voltage drop in the armature resistance is $\mathbf{I_aR_a}$.It is represented by phasor AC.The voltage drop in the synchronous reactance is $\mathbf{I_aX_s}$.It is represented by CD .It leads the current $\mathbf{I_a}$ by 90^0 and, therefore,CD is drawn in a direction perpendicular to OB. The total voltage drop in the synchronous impedance is the phasor sum of $\mathbf{I_aR_a}$ & $\mathbf{I_aX_s}$ .It is represented by AD.The phasor OD represents $\mathbf{E_a}$.

The magnitude $\mathbf{E_a}$ can be found from the right-angled triangle OGD

$$OD^2 = OG^2 + GD^2 = (OF + FG)^2 + (GC + CD)^2$$

$$E_a^2 = (V\,cos\Phi + I_aR_a\,)^2 + (V\,sin\Phi + I_aX_s\,)^2$$

$$E_a = \sqrt{(V\,cos\Phi + I_aR_a\,)^2 + (V\,sin\Phi + I_aX_s\,)^2}$$

……………………………………..1.26

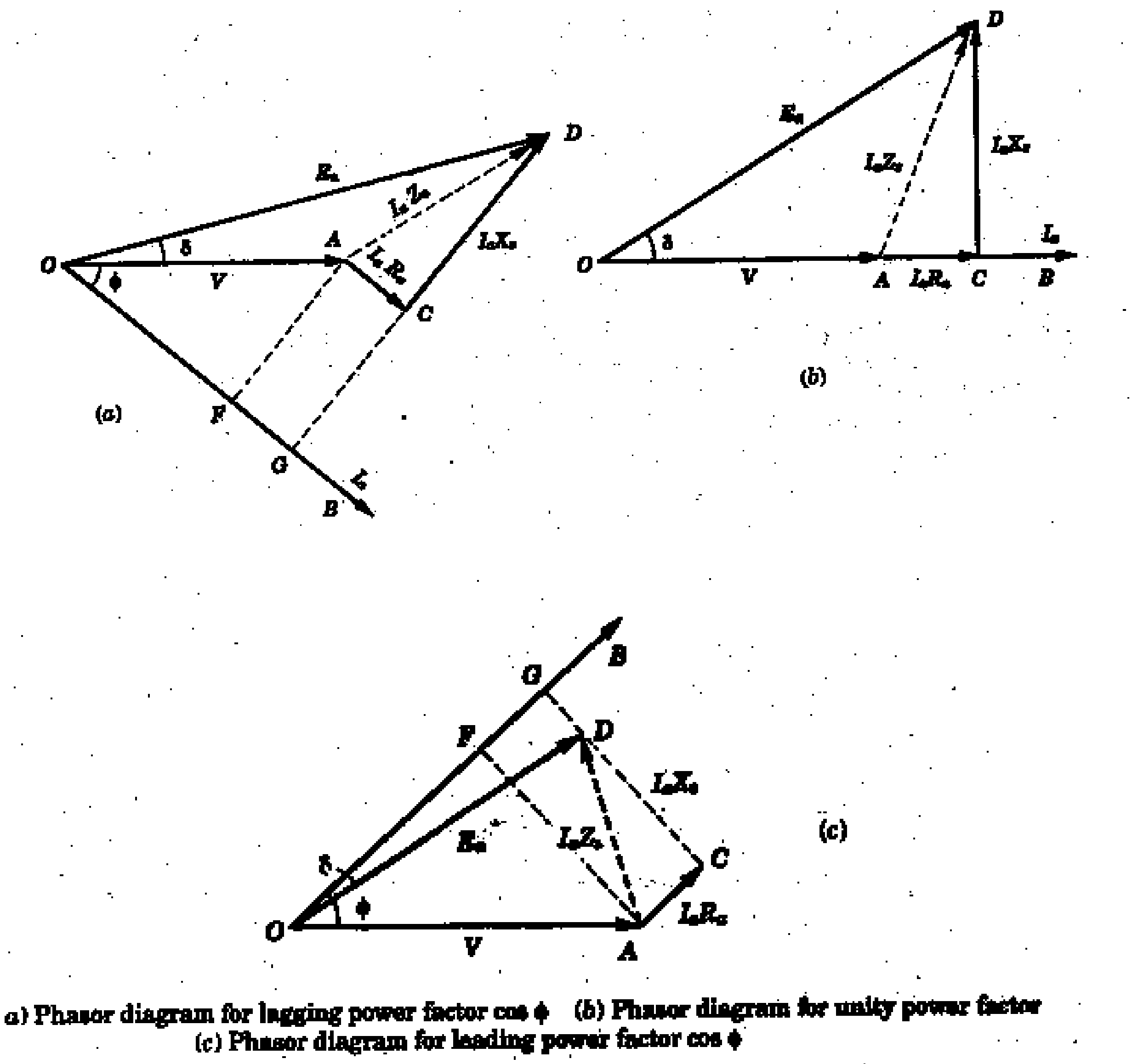


Fig.1.9

b) Unity power factor Cos Φ

Figure 1.9(b) shows the phasor diagram for unity power factor load.

From the right-angled triangle OCD

$$OD^2 = OC^2 + CD^2 = (OA + AC)^2 + (CD)^2$$

$$E_a^2 = (V + I_a R_a)^2 + (I_a X_s)^2$$

$$E_a = \sqrt{(V + I_a R_a)^2 + (I_a X_s)^2} \quad \ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots 1.27$$

c) Leading power factor Cos Φ

Figure 1.9(c) shows the phasor diagram for leading power factor load.

From the right-angled triangle OGD

$$OD^2 = OG^2 + GD^2 = (OF + FG)^2 + (GC - CD)^2$$

$$E_a^2 = (V\,cos\Phi + I_a R_a)^2 + (V\,sin\Phi - I_a X_s)^2$$

$$E_a = \sqrt{(V\,cos\Phi + I_a R_a)^2 + (V\,sin\Phi - I_a X_s)^2}$$
…………………………………..1.28

The angle δ between E_a & V is called the **power angle** or **Torque angle** of the machine. It varies with load and is a measure of air-gap power developed in the machine.

VOLTAGE REGULATION

The voltage regulation of as synchronous generator is defined as the change in terminal voltage from no-load to full-load divided by the full-load voltage when the speed and field current remaining constant.

It is expressed as a fraction or a percentage of full-load terminal voltage .It can be written as

Per unit voltage regulation = $\frac{E_0 - V}{V}$

Percentage voltage regulation == = $\frac{E_0 - V}{V}$ x 100

Where E_0 = magnitude of generated voltage per phase

V = magnitude of rated terminal voltage per phase

The voltage regulation depends upon the power factor of the load. For unity and lagging power factors, there is always a voltage drop with the increase of load, but for a certain leading p.f. the full-load regulation is zero. In this case the terminal voltage is same for both full-load and no-load conditions. At lower leading power factors the voltage rises with the increase of load,and the regulation is negative.

DETERMINATION OF VOLTAGE REGULATION

The KVA ratings of commercial alternators are very high(.Example 500MVA). It is neither convenient nor practicable to determine the voltage regulation by direct loading. There are several indirect methods of determining the voltage regulation of an alternator. These methods require only a small amount of power as compared to the power required for a direct loading method.

Three such methods are

(i) **Synchronous impedance method**

(ii) **Ampere-Turn method**

(iii) **Zero power factor or Potier method**

For the synchronous impedance method the following tests are conducted .

MEASUREMENT OF SYNCHRONOUS IMPEDANCE

The following tests are performed on an alternator to know itsb performance.

a) DC resistance test
b) Open-circuit test
c) Short-circuit Test

DC RESISTANCE TEST

Assume that the alternator is star connected with d.c. field winding open(Fig. 1.10) , measure the d.c. resistance between each pair of terminals either by using ammeter-voltmeter method or by using Wheat stone's bridge. The average of three sets of resistance values R_t is taken. This value of R_t is divided by 2 to get the d.c. resistance(ohmic resistance) per phase.The alternator should be at rest. Since the effective a.c. resistance is larger than d.c. resistance due to skin effect, therefore, effective a.c. resistance per phase is obtained by multiplying the d.c. resistance by a factor of 1.2 to 1.75 depending on the size of the machine.

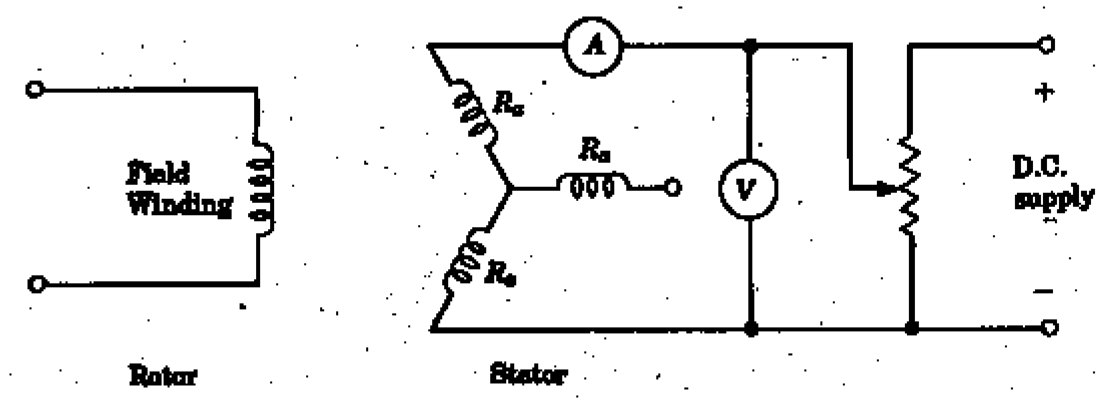


Fig.1.10 D.C. resistance test on an alternator.

Open –circuit Test

The alternator is run at rated synchronous speed and the load terminals are kept open Fig.1.11. That is, all the loads are disconnected .The field current is set to zero.

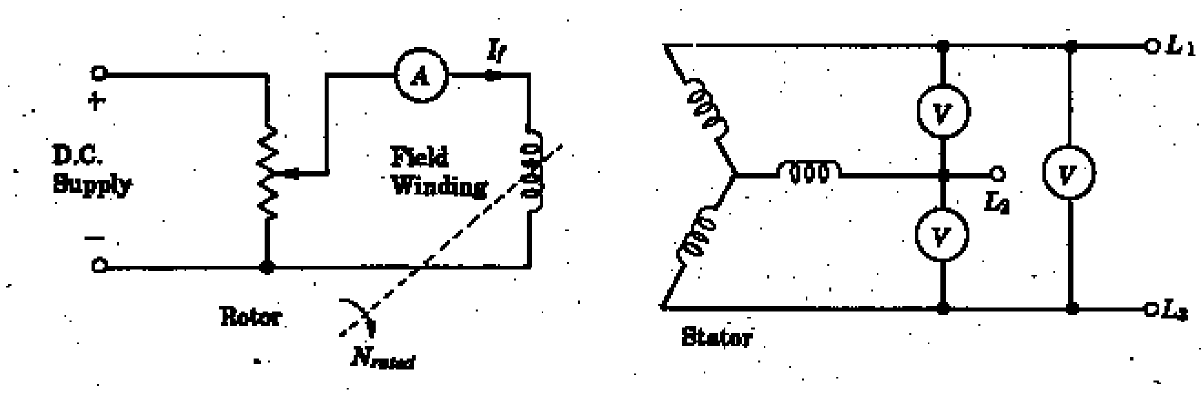


Fig.1.11 **Open-circuit test on an alternator.**

Then the field current is gradually increased in steps, and the terminal voltage E_t is measured at each step.The excitation current may be increased to get 25% more than rated voltage of the alternator.A graph is plotted between the open-circuit. phase voltage E_p $(=E_t/\sqrt{3})$ and field current I_f .The characteristics curve so obtained is called open-circuit characteristics(O.C.C.).It takes the shape of a normal magnetization curve.The extension of the linear portion of an O.C.C is called the air-gap line of the characteristics. The O.C.C. and the air-gap line are shown in figure 1.12.

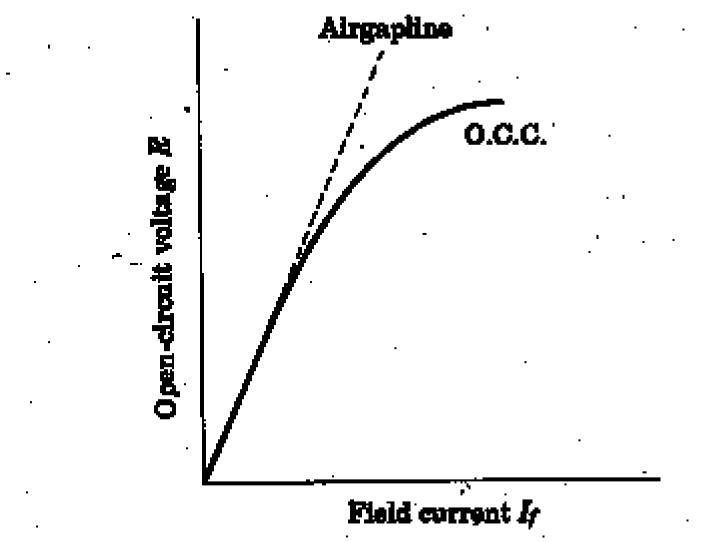


Fig.1.12 The O.C.C. of an alternator

Short-Circuit Test

The armature terminals are shorted through three ammeters (Fig. 1.13) . Care should be taken in performing this test, and the field current should first be decreased to zero before starting the alternator. Each ammeter should have a range greater than the rated full-load value. The alternator is then run at synchronous speed. Then the field current is gradually increased in steps , and the armature current is measured at each step. The field current may be increased to get armature currents upto 150% of the rated value. The field current I_f and the average of three ammeter readings at each step is taken. A graph is plotted between the armature current I_a and the field current I_f .The characteristics so obtained is called *short-circuit characteristics* (SCC).This characteristic is a straight line as shown in figure 1.14.

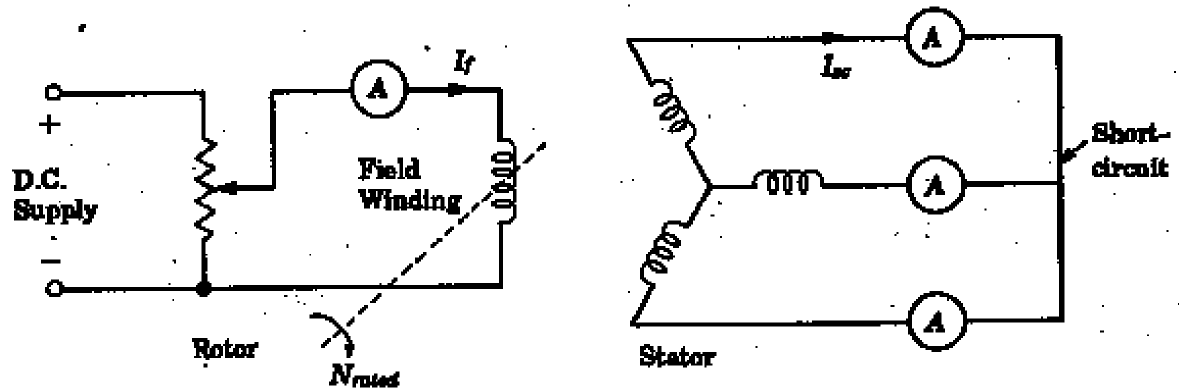


Fig.1.13

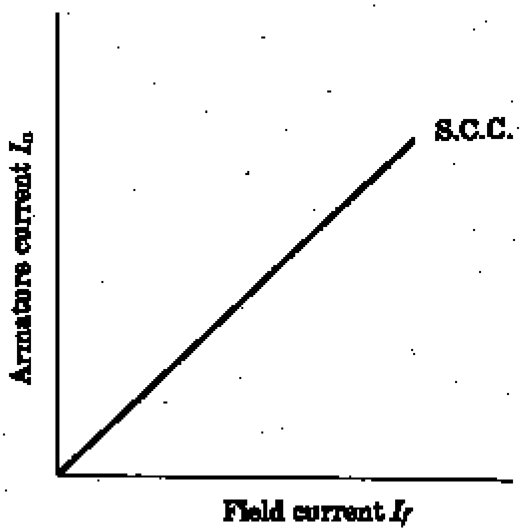


Fig.1.14.The S.C.C of an alternator

Calulation of Zs

The open-circuit characteristic (O.C.C.) and short-circuit characteristic(S.C.C) are drawn on the same curve sheet. Determine the value of I_{sc} at the field current that gives the rated alternator voltage per phase. The synchronous impedance Z_s will then be equal to the Open-Circuit voltage divided by the short circuit current at that field current which gives the rated e.m.f. per phase.

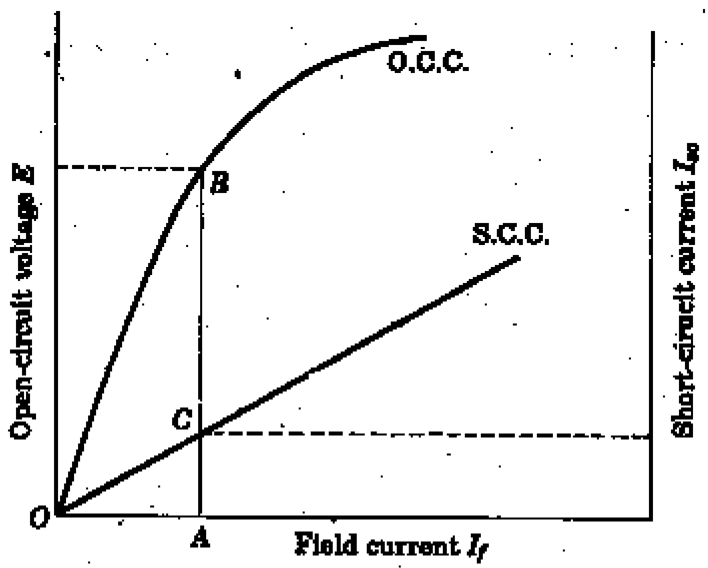


Fig.1.15

Zs = $\frac{open\ circuit\ voltage\ per\ phase}{short-circuit\ armature\ current}$ for the same value of field current.

Now, knowing the Synchronous impedance and Armature resistance we can find out the Synchronous reactance

synchronous reactance $X_s = \sqrt{Z_s^2 - R_a^2}$

Once we know R_a and X_s,the phasor diagram can be drawn for any load and any P.F.. Fig 1.16. shows the phasor diagram for the usual case of inductive load ; the load power factor being $\cos\emptyset$ lagging. Note that in drawing the phasor diagram,current I_a has been taken as the reference phasor. The I_aR_a drop is in phase with I_a while I_aX_s drop leads I_a by 90^0.The phasor sum of V, I_aR_a and I_aX_s gives the no-load e.m.f. E_0 .

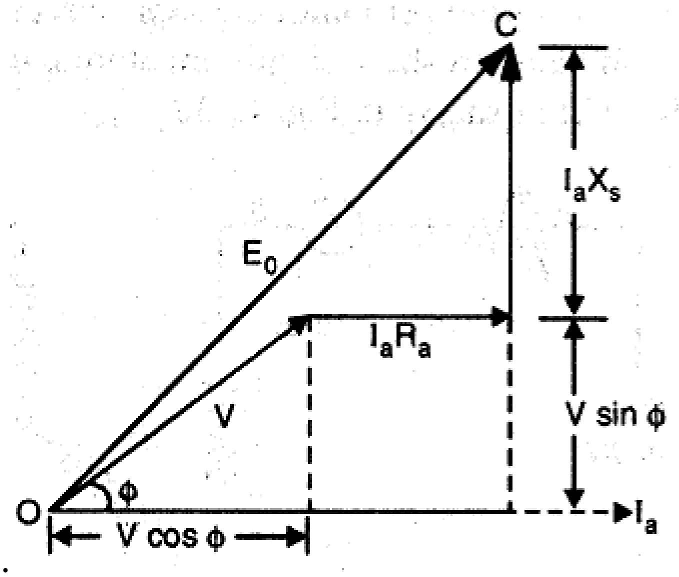


Fig.1.16

Now,

$E_0 = \sqrt{(OB)^2 + (BC)^2}$

OB= V $\cos\emptyset$ + I_aR_a and BC= V $\sin\emptyset$+ I_aX_s

$\therefore$ $E_0 = \sqrt{(V\cos\emptyset + I_aR_a)^2 + (V\sin\emptyset + I_aX_s)^2}$

$\therefore$ % voltage regulation $= \frac{E_0 - V}{V} \times 100$

Drawback : This method is easy but it gives approximate results .The reason is simple. The combined effect of X_L(armature leakage reactance) and X_{AR}(reactance of armature reaction) is measured on short-circuit.Since the current in this condition is almost lagging 90^0 ,the armature reaction will provide its worst demagnetizing effect. It follows that under any normal operation at,say 0.8 or 0.9 lagging power factors will produce error in calculations. This method gives a value higher than the value obtained from an actual load test.For this reason, it is called *pessimistic method.*

Example 1.3

A 500 v , 50 kVA single-phase alternator has an effective armature resistance of 0.2 Ω . An excitation current of 10 A produces 200 A armature current on short-circuit and an e.m.f. of 450 volt on open circuit.Calculate the synchronous reactance.

Solution :

$$Zs = \frac{open\ circuit\ voltage\ per\ phase}{short-circuit\ armature\ current} = \frac{450}{200} = 2.25\ \Omega$$

$$Xs = \sqrt{Z_s^2 - R_a^2} = \sqrt{2.25^2 - 0.2^2} = 2.241\Omega$$

Example 1.4

A 3-phase 2300 V ,50 Hz,1500 kVA star-connected alternator has a resistance between each pair of nterminals as measured by direct current is 0.16 Ω. Assume that the effective resistance is 1.5 times the ohmic resistance.A field current of 70A produces a short-circuit current equal to full-load current of 376 A in each line.The same field current produces an e.m.f. of 700 volt on open circuit.Calculate the synchronous reactance of the machine and its full-load regulation at 0.8 power factor Lagging

Solution :

$$Zs = \frac{open\ circuit\ voltage\ per\ phase}{short-circuit\ armature\ current} = \frac{700/\sqrt{3}}{376} = 1.075\ \Omega$$

Ohmic resistance per phase = $\frac{0.16}{2}$ = 0.08 Ω

Effective resistance per phase = R_a = 1.5 x 0.08 = 0.12 Ω

Synchronous reactance = $X_s = \sqrt{Z_s^2 - R_a^2} = \sqrt{1.075^2 - 0.12^2}$ = 1.068Ω

$S_{3\emptyset} = \sqrt{3}\ V_L I_L$; => 1500 x $10^3 = \sqrt{3} \times 2300\ I_L$; => I_L = 376 A

Rated voltage per phase = $V_p = 2300/\sqrt{3}$ = 1328 V

Phase current $I_{ap} = I_L$ =376 A

$E_p = V_p + I_{ap} Z_s$

Let V_p be taken as reference phasor :

$V_p = V_p L\ 0^0$ = 1328 $L0^0$ volts =1328 + j 0 volts

$I_{ap} = I_{ap} L \cos^{-1} 0.8$ =376 L-36.87^0 A

$Z_s = R_s + jX_s$ = 0.12+j 1.068 =1.075L$83.59^0 \Omega$

E_p= 1328 + j 0 +(376 L-36.87^0)(1.075L83.59^0) = 1328+404.2L46.72^0

= 1328+277.1+j 294.26 = 1605.1+j 294.26 =1631L10.39^0 volt

Percentage Regulation = $\frac{E_P - V_P}{V_P}$ x 100 = 22.8 %

AMPERE-TURN METHOD

This method of finding voltage regulation considers the opposite view to the synchronous impedance method. It assumes the *armature leakage reactance to be additional armature reaction.* Neglecting armature resistance(always small),this method assumes that change in terminal p.d. on load is due entirely to armature reaction. The same two tests(Open and short circuit test)are required as for synchronous reactance determination ; the interpretation of the results only is different. Under short-circuit, the current lags by 90^0(R_a considered zero) and the power factor is zero.Hence,the armature reaction is entirely demagnetizing. Since,the terminal p.d. is zero,all the field AT(ampere-turns)are neutralized by armature AT produced by the short circuit armature current.

(i) Suppose the alternator is supplying full-load current at normal voltage V(i.e.,operating load voltage)and zero p.f. lagging. Then d.c. field AT required will be those needed to produce normal voltage V(or if R_a is to be taken into account,then V+ $I_a R_a \cos\emptyset$) on no-load plus those to overcome the armature reaction.

Let AO =field AT required to produce the normal voltage V (or V+$I_a R_a \cos\emptyset$) at no-load.

OB_1=field AT required to neutralize the armature reaction

Then total field AT required are the phasor sum of AO+OB_1 (Fig.1.17(i))

total field AT, AB_1 = AO+OB_1 ………phasor sum

The AO can be found from O.C.C. and OB_1 can be determined from S.C.C.. Note that the use of a d.c. quantity(field AT)as a phasor is perfectly valid in this case because the d.c. field is rotating at the same speed as the a.c. phasors i.e. $\omega = 2\pi f$.

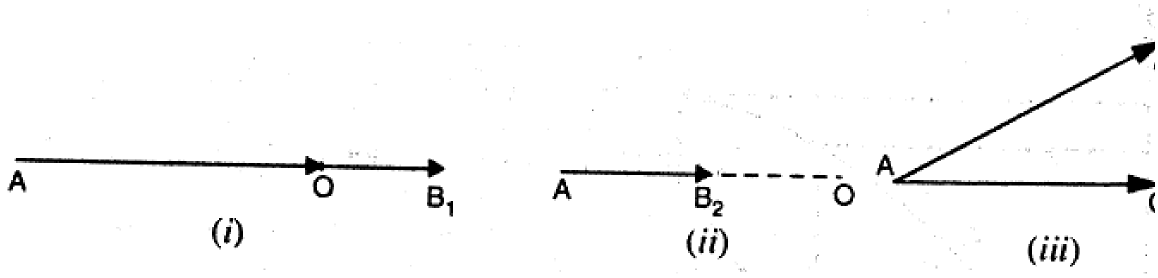


Fig 1.17

(ii) For a full-load current of zero p.f. leading, the armature AT are unchanged. Since they aid the main field, less field AT are required to produce the given e.m.f.

So, the total field AT, AB_2 = AO-B_2O ……………………………phasor difference

Where B_2O =field AT required to neutralize armature reaction

This is illustrated in Fig 1.17(ii).Note that again AO is determined from O.C.C. and B_2O from S.C.C.

(iii)Between zero lagging and zero leading power factors,the armature m.m.f. rotates through 180^0. At unity p.f. ,armature reaction is cross-magnetising only.Therefore, OB_3 is drawn perpendicular to AO(Fig.1.17(iii)).Now AB_3 shows the required AT in magnitude and direction.

General Case .

It may now be discussed the case when the pf has any value between zero (lagging or leading) and inity.If the power-factor is $\cos\emptyset$ lagging,then $\emptyset$ is laid off to the right of the vertical line OB_3 as shown in Fig.1.18(i).The total field AT required are AB_4 i.e.,phasor sum of AO and OB_4.If the power factor is $\cos\emptyset$ leading , then $\emptyset$ is laid off to the left of the vertical line OB_3 as shown in Fig.1.18(ii).The total field AT required are AB_5 i.e.,phasor sum of AO and OB_5.

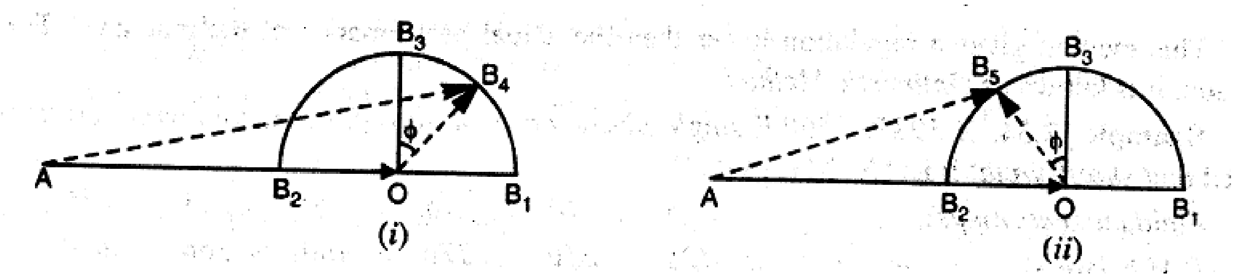


Fig 1.18

Since current $\propto$ AT , it is more convenient to work in terms of field current.Fig. 1.19 shows the current diagram for the usual case fo lagging power factor.Here AO represents the field current required to produce normal voltage V (or $V + I_a R_a \cos\Phi$) on no-load. The phasor OB represents the field current required for producing full-load current on short –circuit.The resultant field current is AB and the phasor sum of AO and OB.Note thatphasor AB represents the field current required for demagnetizing an dto produce voltage V and $I_a R_a \cos\Phi$ drop.(if R_a is taken into account).

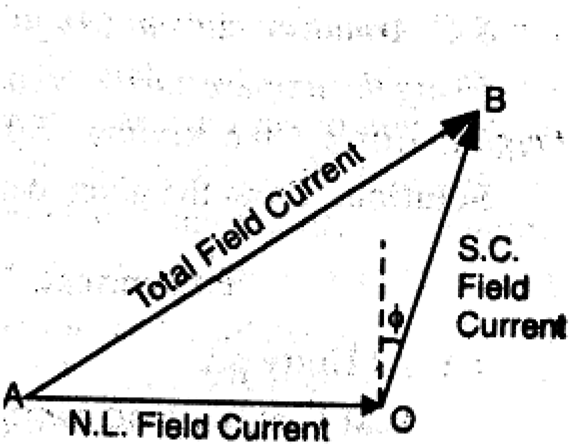


Fig.1.19

PROCEDURE FOR AT METHOD

Suppose the alternator is supplying full-load current I_a at operating voltage Vand p.f. $\cos\Phi$ lagging.The procedure for finding voltage regulation for AT method is as under :

(i) From the O.C.C.,field current OA required to produce the operating load voltage V(or $V + I_a R_a \cos\Phi$) is determined as shown in Fig.1.20.The field current OA is alid off horizontally as shown in Fig.1.21.

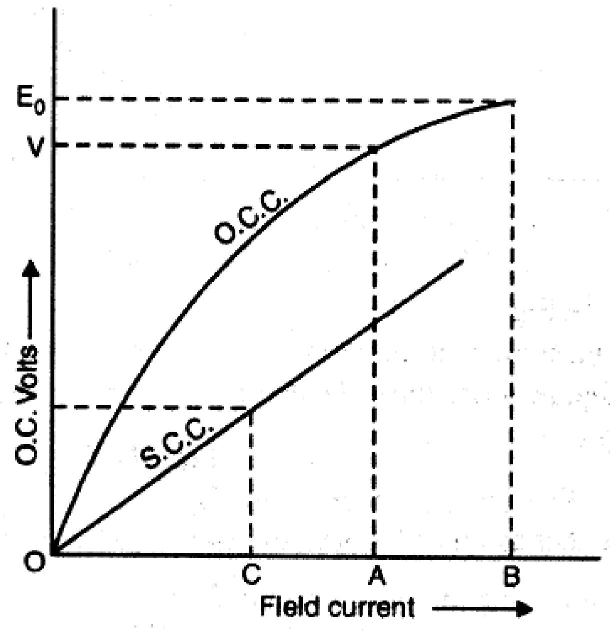


Fig.1.20

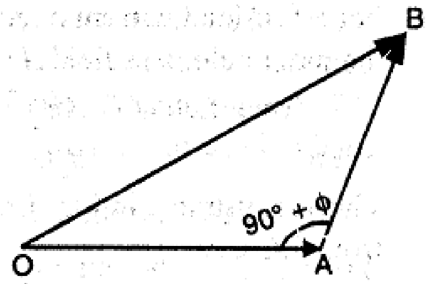


Fig.1.21

(ii) From S.C.C.,the field current OC required for producing full-load current I_a on short-circuit is determined.The phasor AB(=OC) is drawn at an angle of $(90^0 + \emptyset)$ i.e.,Arg(OAB)=$(90^0 + \emptyset)$ as shown in Fig. 1.21.

(iii)The phasor sum of OA and AB gives the total field current OB required.The O.C. voltage $\mathbf{E_0}$ corresponding to field current OB on O.C.C. is the no-load e.m.f.

voltage regulation $=\frac{E_0 - V}{V} \times 100$

This method gives a regulation lower than the actual performance of the machine.For this reason,it is known as *Optimistic method*.

ZERO POWER FACTOR METHOD OR POTIER METHOD

In this method, we separately determine the voltage drop due to armature leakage reactance(=I_aX_l) and voltage drop due to armature reaction(=I_aX_{AR}).Therefore, it gives more accurate results. The Potier method consists of the following steps:

(i) Plotting O.C.C. :

The open-circuit characteristics (O.C.C.) of the alternator is plotted by conducting no-load test on the alternator as explained earlier .The lower part of O.C.C. is practically a straight line and when extended becomes the air-gap line. Therefore, air-gap line represents O.C.C. of the alternator if the reluctance of the iron portion of the magnetic circuit of the machine is neglected as compared to the reluctance of the air-gap.

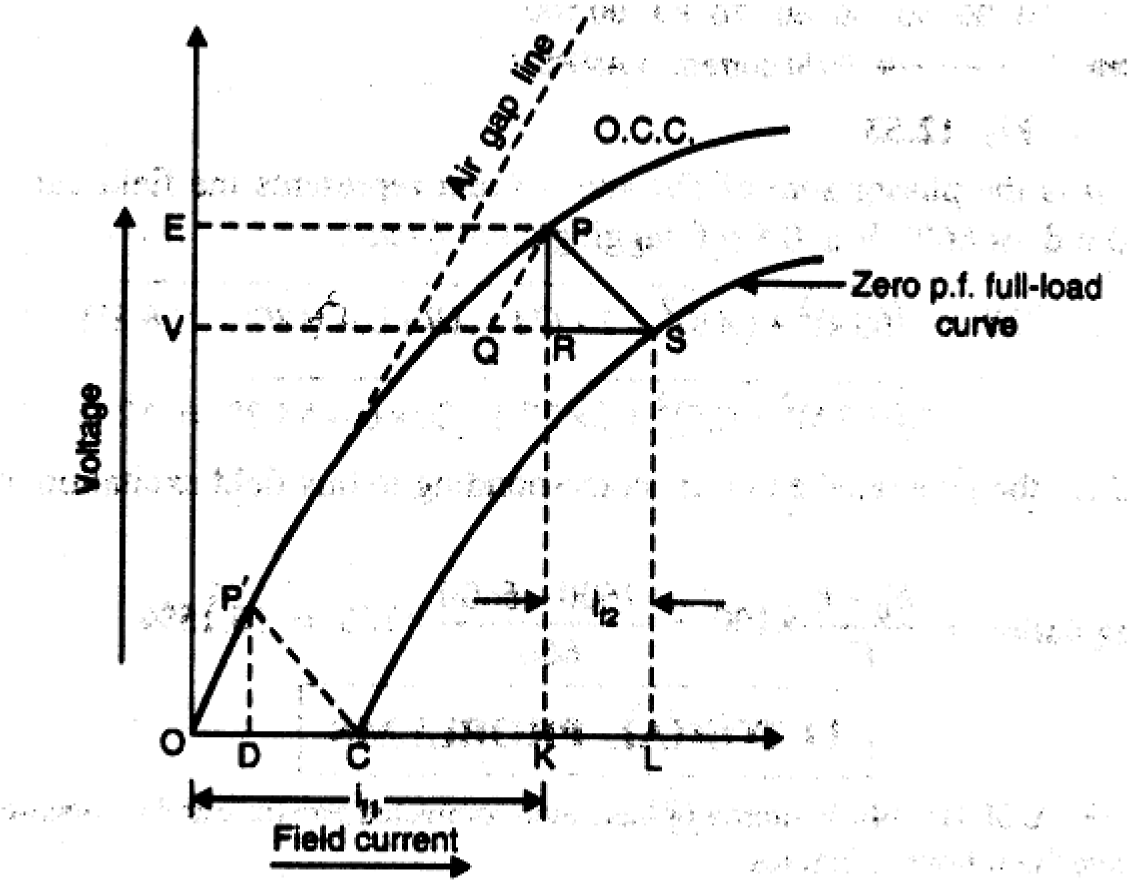


Fig.1.22

(ii) **Plotting zero p.f. (lagging) full-load curve** :

This is the curve between the terminal voltage and field current when the alternator is delivering its full rated current to a zero power factor(lagging) load. The test is carried out by running the alternator at synchronous speed and connecting a purely inductive 3-phase load to its terminals. The load is varied in steps and at each step, the field current is adjusted so that the armature current is equal to its rated value. There is no need to plot the full curve. Only two points S &C (see fig.1.22) are sufficient. The point S corresponds to a field current which gives the rated terminal voltage while the zero p.f. load is adjusted to draw the rated armature current. The point C corresponds to the short-circuit conditions on the alternator(i.e. terminal voltage =0) with the field current adjusted to give rated armature current. Since the armature resistance is negligible, the short-circuit current lags behind the resultant induced e.m.f. by 90^0.Therefore,point C constitutes a point on the zero p.f. curve.

(iii) **Costructing potier triangle** :

Referring to Fig. 1.22 OC is field current producing full-load armature current on short-circuit(the current lags by 90^0). Therefore, the field current OC must be sufficient to counter the demagnetising effect of armature reaction and armature leakage reactance to drop at full-load. From S, draw SQ equal to and parallel to OC. From point Q, draw a line QP parallel to air-gap line. This line meets O.C.C at point P. From point P, draw PR perpendicular to QS and meeting it at point R. The triangle PRS is known as POTIER TRIANGLE. The following information is conveyed by the potier triangle :

a) The length PR represents the armature leakage reactance drop(I_aX_l).
b) The length RS(=KL=i_{f2}) represents the field current to overcome the demagnetising effect of armature reaction at full-load.The length RQ represents the field current necessary to balance the armature leakage reactance drop(I_aX_l).
c) The Potier triangle is the same for a given armature current and hence can be transferred anywhere to fit between the two characteristics. The potier triangle PRS is transferred to the baseline as triangle P'DC and is identical to Potier triangle PRS.

PHASOR DIAGRAM : Suppose the full-load terminal voltage of the alternator is V. Let the load p.f.be $\cos\emptyset$ lagging.The phasor diagram for this condition is shown in Fig.1.23

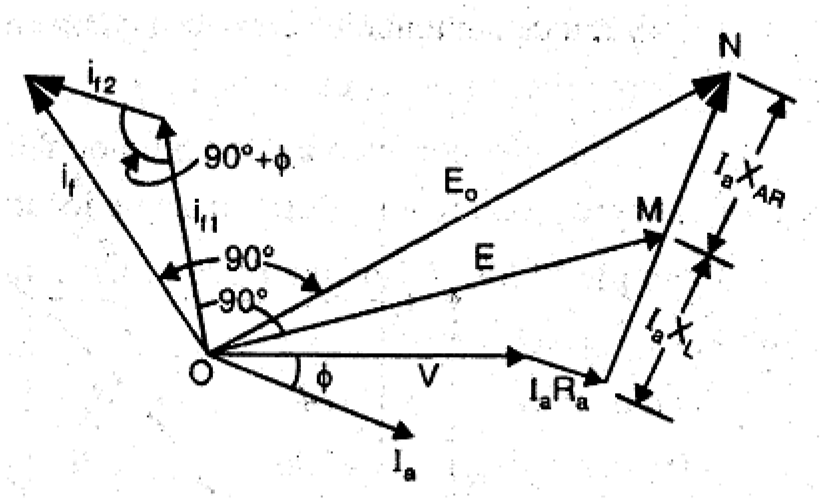


Fig.1.23

Here the voltage V is taken as the reference phasor and current I_a lags behind V by $\emptyset$. The I_aR_a drop is drawn parallel to the current phasor and I_aX_L drop is drawn perpendicular to it.The phasor OM represents the induced e.m.f E(=phasor sum of V, I_aR_a and I_aX_L).From O.C.C. in Fig.1.22,the field current corresponding to induced e.m.f. E is i_{f1} (=OK).The field current i_{f1} is drawn 90^0 ahead of E. The current phasor i_{f2} (=RS=DC in Fig.1.22)represents the field current necessary to overcome the demagnetising effect of armature reaction at full-load and is drawn parallel to the current phasor I_a.The phasor sumof i_{f1} & i_{f2} gives the total field current i_f required to produce E_0.The phasor E_0(=ON) lags behind i_f by 90^0.Note that phasor MN represents the voltage drop $I_a X_{AR}$ due to armature reaction.

$$\text{Percentage voltage regulation} = \frac{E_0 - V}{V} \times 100$$

PROCEDURE FOR POTIER METHOD

For solving problems relating to Potier method, the following procedure is adopted:

(i) Suppose the terminal voltage per phase is V.

(ii) Find the armature leakage reactance drop($= I_a X_L$) from the potier triangle.

(iii) Add $I_a X_L$)(and $I_a R_a$ if given) vectorially to V to get E.

(iv) From the O.C.C, find the field current required to produce E. Let it be i_{f1} .

(v) From the potier triangle, find the field current necessary for balancing armature reaction. Let it be i_{f2}.

(vi) Find the phasor sum of i_{f1} & i_{f2} to get the resultant field current i_f .

(vii) From the O.C.C. , find the e.m.f. corresponding to field current i_f . This gives us E_0 .Therefore, we can find the voltage regulation.

LIMITATIONS OF POTIER METHOD

The Potier method has the following drawbacks :

(a) The Potier triangle is based on the assumption that the armature leakage reactance is constant and the O.C.C. of the alternator is the same under load as it is under open-circuit conditions. However ,this is not correct. Therefore, the value of armature leakage reactance determined by the potier triangle method is not absolutely accurate. In order to distinguish the value of leakage reactance thus determined from the absolutely correct value, the value of leakage reactance determined by the potier triangle method is sometimes called potier leakage reactance.

(b) In the potier triangle method, some error is also introduced due to the fact that it is not possible to obtain exactly zero P.F. for the zero power factor load test.

(c) A greater amount of error is exhibited by machines of salient pole construction than the non-salient pole construction. Best results are obtained by determining the value of armature leakage reactance and the m.m.f. of armature reaction from a potier triangle located as high as possible on the test curves.

PARALLEL OPERATION OF ALTERNATORS

It is rare to find a 3-phase alternator supplying its own load independently except under test conditions. In practice, a very large number of 3-phase alternators operate in parallel because the various power stations are interconnected through the national grid. Therefore, the output of any single alternator is small compared with the total inter- connected capacity. For example, the total capacity of the interconnected system may be over 40,000MW while the capacity of the biggest single alternator may be 500MW.For this reason, the performance of a single alternator is unlikely to affect appreciably the voltage and frequency of the whole system .An alternator connected to such a system is said to be connected to *infinite bus-bars*. The outstanding electrical characteristics of such bus-bars are that they are constant – voltage ,constant-frequency bus-bars.

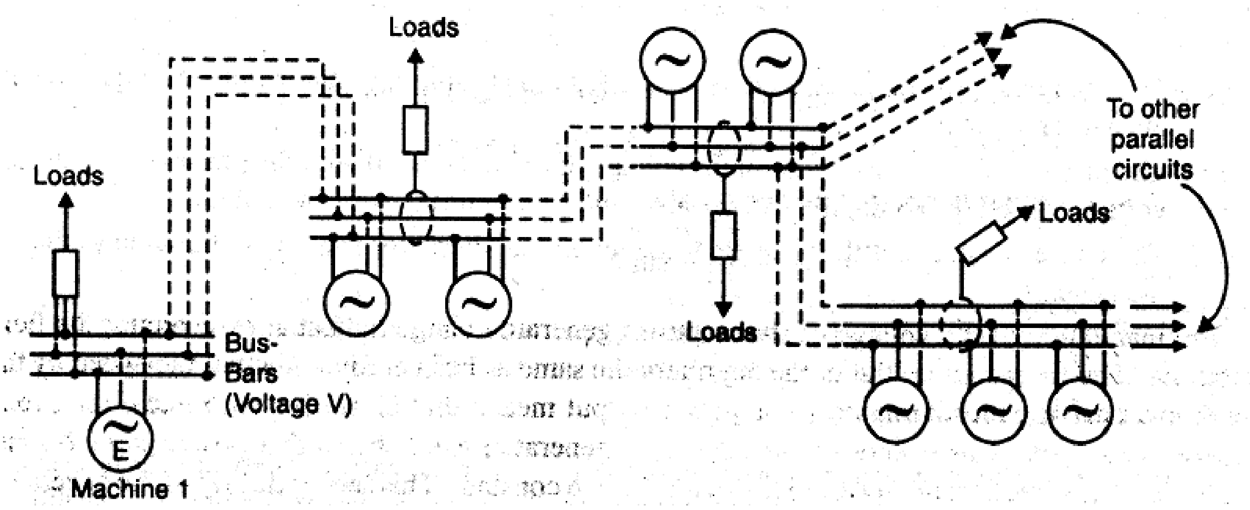


Fig.1.24

Fig. 1.24 Shows a typical infinite bus system .Loads are tapped from the infinite bus at various load centers. The alternators may be connected to or disconnected from the infinite bus, depending on the power demand on the system. If an alternator is connected to infinite bus-bars, no matter what power is delivered by the in-coming alternator, the voltage and the frequency of the system remain the same. The operation of connecting an alternator to the infinite bus-bars is known as parallel with the infinite bus-bars. It may be noted that before an alternator is connected to an infinite bus-bars, certain conditions must be satisfied.

ADVANTAGES of PARALLEL OPERTIONS OF ALTERNATORS

The following are the advantages of operating alternators in parallel.

(i) Continuity of service : The continuity of service is one of the important requirements of any electrical apparatus. If one alternator fails, the continuity of supply can be maintained through the other healthy units. This will ensure un-interrupted supply to the consumers.

(ii)Efficiency :The load on the power system varies during the whole day ; being minimum during the late night hours. Since, alternators operate most efficiently when delivering full-load ,units can be added or put-off depending upon the load requirement. This permits the efficient operation of power system.

(iii)Maintainance and Repair : It is often desirable to carry out routine maintenance and repair of one or more units. For this purpose , the desired unit /units can be shut down and the continuity of supply is maintained through the other units.

(iv)Load growth: The load demand is increasing due to the increase in use of electrical energy. The load growth can be met by adding more units without disturbing the original installation .

CONDITIONS FOR PARALLELING ALTERNATOR WITH INFINITE BUSBARS

The proper method of connecting an alternator to the infinite bus-bars is called synchronising. A stationary alternator must not be connected to live bus-bars. It is because the induced e.m.f. is zero at stand-still and a short-circuit will result .In order to connect an alternator safely to the infinite bus-bars , the following conditions are met.

(i) The terminal voltage (r.m.s. value) of the incoming alternator must be the same as bus-bars voltage.

(ii) The frequency of the generated voltage of the incoming alternator must be equal to the bus-bars frequency.

(iii) The phase of the incoming alternator voltage must be identical with the phase of the bus-bars voltage. In other words, the two voltages must be in-phase with each other.

(iv) The phase sequence of the voltage of the incoming alternator should be same as that of the bus-bars.

The magnitude of the voltage of the incoming alternator can be adjusted by changing its field excitation. The frequency of the incoming alternator can be changed by adjusting the speed of the prime-mover driving the alternator.

Condition(i) is indicated by a voltmeter, condition (ii) & (ii) are indicated by synchronizing lamps or a synchroscope. The condition (iv) is indicated by a phase sequence indicator.

METHODS OF SYNCHRONISATION

The method of connecting an incoming alternator safely to the live busbars is called synchronising. The quality of voltage between the incoming alternators and the busbars can be easily checked by a voltmeter. The phase sequence of the alternator and the busbars can be checked by a phase sequence indicator. Differences in frequency and phase of the voltages of the incoming alternators and busbars can be checked by one of the following two methods.

1. **By three lamps(one dark,two bright) method**.

2. **By synchroscope**

Three-Lamp method : In this method of synchronising ,three lamps $L_1, L_2\ and\ L_3$ are connected as shown inFig1.25.The lamp L_1 is straight connected between the corresponding phases (R_1&R_2) and the other two are cross-connected between the other two phases. Thus, lamp L_2 is connected between Y_1 & B_2 and lamp L_3 between B_1 & Y_2 .When the frequency & phase of the voltage of the incoming alternator is the same as that of the busbars, the straight connected lamp L_1 will be dark while cross-connected lamp L_2 & L_3 will be equally bright. At this instant ,the synchronisation is perfect and the switch of the incoming alternator can be closed to connect it to the busbars. In Fig.1.25 Phasors R_1, Y_1 & B_1 represent the busbar voltages and phasors R_2, Y_2 & B_2 represent the voltages of the incoming alternator. At the instant when R_1 is inphase with R_2 ,voltage across lamp L_1 is zero and voltages across lamps

L_2 &L_3 are equal. Therefore ,lamp L_1 is dark while lamps L_2 &L_3 will be equally bright .At this instant, the switch of the incoming alternator can be closed.Thus,incoming alternator gets connected in parallel with the busbars.

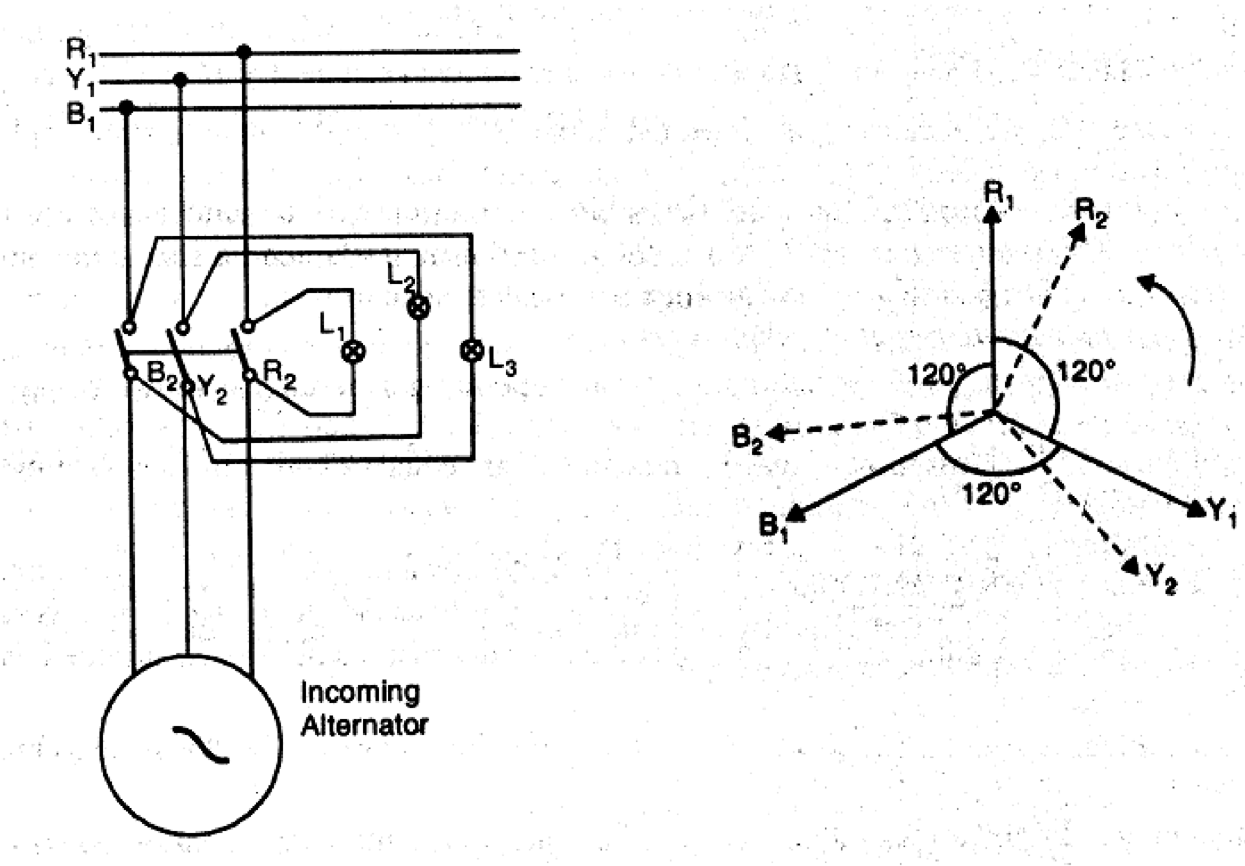


Fig.1.25

Synchroscope : A synchroscope is an instrument that indicates by means of a revolving pointer the phase difference and the frequency difference between the voltages of the incoming alternator and the busbars(Fig. 1.26). It is essentially a small motor ,the field being supplied from the busbars through a potential transformer and the rotor from the incoming alternator. Pointer is attached to the rotor. When the incoming alternator is running fast ,(i.e.the frequency of the incoming alternator is higher than that of the busbars),the rotor and hence pointer moves in the clock-wise direction. When the incoming alternator is running slow that is ,(frequency of the incoming alternator is lower than that of the busbars),the pointer moves in anti-clockwise direction.When the frequency of the incoming alternator is equal to that of the busbars , no torque acts on the rotor and the pointer points vertically upwards ("12 O' Clock").It indicates the correct instant for connecting the incoming alternator to the busbars. The synchroscope method is superior to the Lamp method because it not only gives a positive indication of the time to close the switch but also indicates the adjustments to be made should there be a difference between the frequencies of the incoming alternator and the busbars.

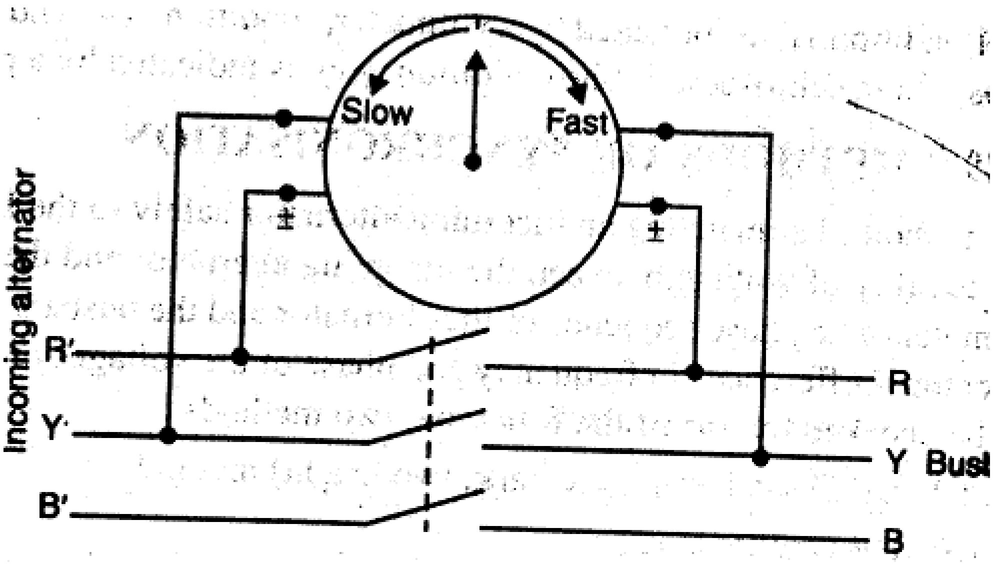


Fig.1.26

Synchronising action :

When two or more alternators have been connected in parallel ,they will remain in stable operation under all normal conditions i.e., voltage ,frequency, speed and phase equality will continue .In other words, once synchronised properly, the alternators will continue to run in synchronism under all normal conditions. If one alternator tries to fall out of synchronism ,it is immediately counteracted by the production of a synchronising torque which brings it back to synchronism. This automatic action is called the synchronising action of the alternators.

Consider two similar single-phase alternators 1 &2 operating in parallel as shown in Fig. 1.27 (i).For simplicity, let us assume that the alternators are at no-load. When in exact synchronism , magnitudes of the small e.m.f. s $\mathbf{E_1}$ (Machine 1) & $\mathbf{E_2}$ (machine 2) are equal. These e.m.f.s are acting in the same direction with respect to the external circuit as shown in Fig.1.27(ii).But in relation to each other, these e.m.f.s are in phase opposition i.e. ,if we trace the closed circuit formed by the two alternators we find that the e.m.f.s oppose each other as shown in Fig 1.27(iii).When the alternators are in exact synchronism, $\mathbf{E_1}$ & $\mathbf{E_2}$ are in exact phase opposition.Since, $\mathbf{E_1} = \mathbf{E_2}$ in magnitude, no current flows in the closed circuit formed by the two alternators.

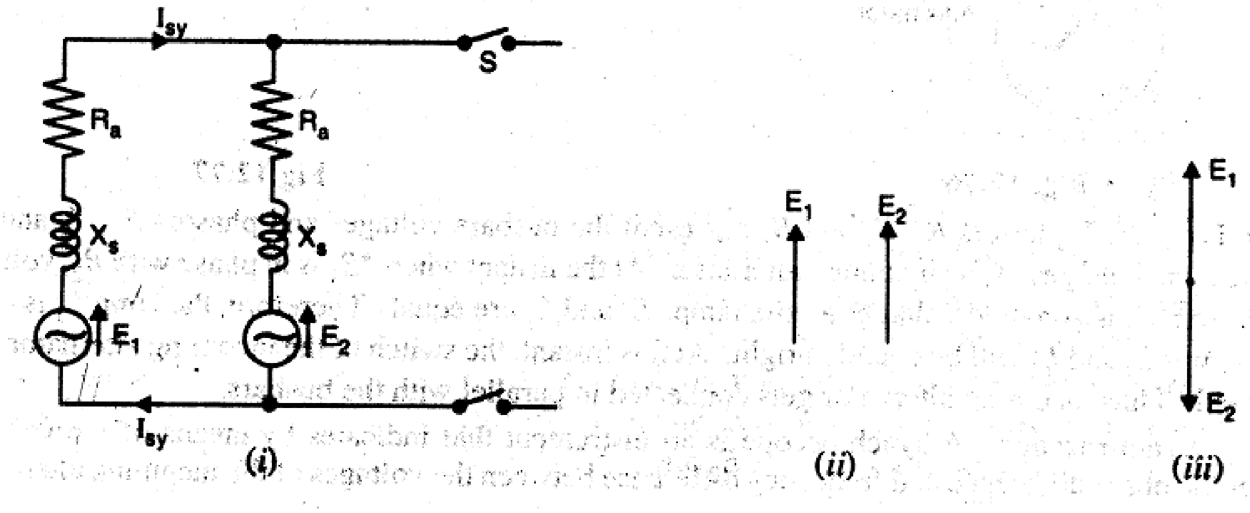


Fig.1.27

If one alternator drops out of synchronism, there is an automatic action to re-establish synchronism.

Let us discuss this point.

(i) **Effect of speed change** : Suppose, due to any reason ,the speed of machine2 falls .Then emf E_2 will fall back a phase angle of α electrical degrees as shown in Fig 1.28.(though still $E_1 = E_2$).There will be resultant e.m.f. E_r in the closed circuit formed by the two alternators.The emf E_r will circulate current (known as synchronising current I_{sy}) in this closed circuit.

Synchronising current $I_{sy} = \frac{E_r}{2Z_s}$

The current I_{sy} lags behind E_r by an angle θ given by :

$$\tan\theta = \frac{2X_s}{2R_s} = \frac{X_s}{R_s}$$

R_a = armature resistance of each alternator

X_s = synchronous reactance of each alternator

Z_s = synchronous impedance of each alternator

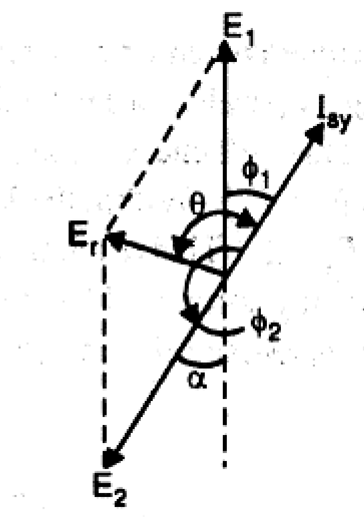


Fig.1.28

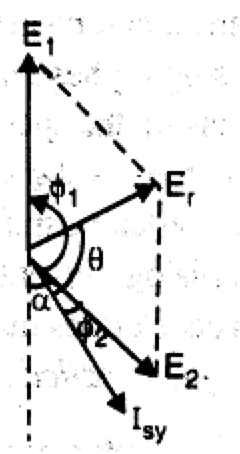


Fig.1.29

Since, R_a is very small as compared to X_s , θ is nearly 90^0 so that the current I_{sy} is almost in phase with E_1 & in phase opposition to E_2 .This means that machine 1 is generating and machine 2 is motoring . consequently, the machine 1 tends to slow down and machine 2, by accepting power , tends to accelerate. This restores the status quo i.e., synchronism is re established.

Conversely,if E2 tends to advance in phase,as shown in Fig.1.29 The directions of E_r and I_{sy} are changed such that now machine 2 is generating and machine 2 is motoring .Once again the synchronism is restored.

(ii) **Effect of inequality of e.m.f.s** : The automatic re-establishment of synchronism of two alternators operating in parallel also extends to any changes tending to alter the individual e.m.f. s . When in exact synchronism ,then $E_1=E_2$(magnitude) and they are in exact phase opposition as shown in Fig.1.30(i). Suppose due to any reason , e.m.f. E_1 increases.Then resultant e.m.f. E_r exists in the closed circuit formed by the two alternators. Then $E_r = E_1 - E_2$ and is in-phase with E_1 .The resultant e.m.f. E_r sends synchronising current I_{sy} in the closed circuit.Here again the current I_{sy} almost lags behind E_r by 90^0(as Zs=Xs) as shown in Fig.1.30(ii). Also I_{sy} lags almost 90^0 behind E_1 and leads E_2 almost by 90^0 .The power produced is practically zero ; just enough to overcome copper losses. The current I_{sy} lags behind E_1 and produces a demagnetising armature reaction effect on machine1 .At the same time I_{sy} leads E_2 and produces magnetising armature reaction effect on the machine2. Thus, E_1 tends to fall and E_2 tends to rise.The result is that synchronism is re-established . The converse is true for $E_2 > E_1$ as shown in Fig.1'30(iii)

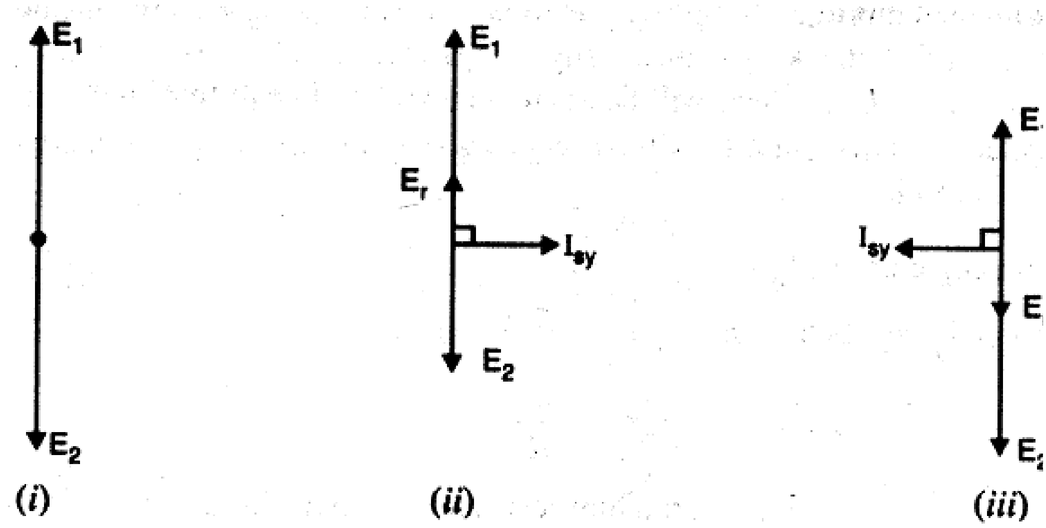
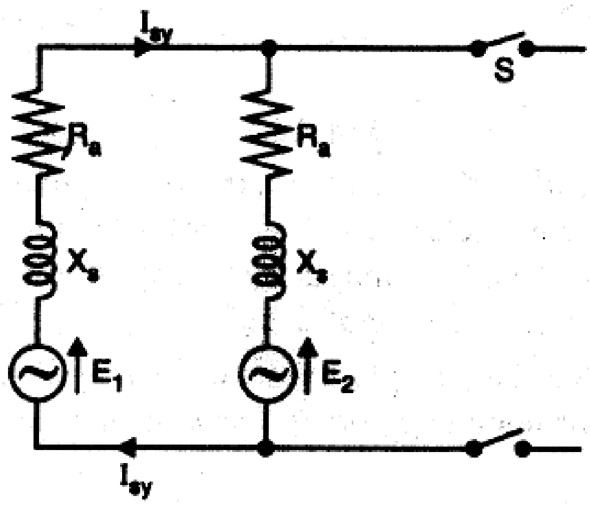


Fig.1.30

SYNCHRONISING POWER :

When two alternators are operating in parallel, each machine has an inherent tendency to remain synchronised . Consider two similar single-phase alternators 1 & 2 operating in parallel at no-load as shown in Fig . suppose ,due to any reason ,the speed of machine 2 decreases. This will cause E_2 to fall back by a phase angle of α electrical degrees as shown in Fig.1.31(though still $E_1=E_2$).Within the local circuit formed by two alternators,the resultant e.m..f E_r is the phasor difference $E_1 - E_2$.This resultant e.m.f. results in the production of synchronising current I_{sy} which sets up synchronising torque. The synchronising torque retards machine 1 and accelerates machine 2 so that synchronism is re-established.The power associated with synchroniosing torque is called synchronizing power.

In Fig.1.32, machine 1 is generating and machine2 is motoring. The power supplied by machine 1 is called **synchronising power**.



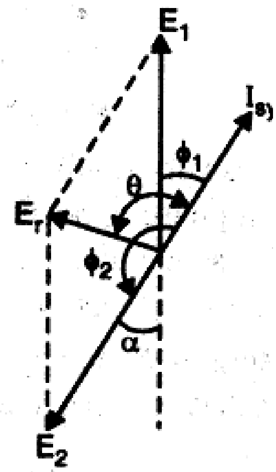


Fig.1.31

Fig.1.32

Referring to Fig.1.32 we have,

Synchronising power , $P_{sy} = E_1 I_{sy} \cos \Phi_1 = E_1 I_{sy} \cos(90^0 - \theta) = E_1 I_{sy} \sin\theta$

$= E_1 I_{sy}$ (As $\theta = 90^0$ so that $\sin\theta = 1$)

The synchronizing power goes to supply power input to machine 2 and the Cu losses in the local circuit of two machines.

$\therefore$ $E_1 I_{sy} = E_2 I_{sy}$ + Cu losses

Resultant e.m.f. , $E_r = 2E\cos\frac{(180^0-\alpha)}{2}$ [As $E_1 = E_2 = E$ (say)]

$= 2\,E\cos(90^0 - \frac{\alpha}{2}) = 2\,E\sin\frac{\alpha}{2} = 2\,E \times \frac{\alpha}{2}$ (As α is small)

$= \alpha E$

Note that in this expression, α is in electrical radians.

synchronising current $I_{sy} = \frac{E_r}{2X_s} = \frac{\alpha E}{2X_s}$ ……………………R_a of both machines is negligible

Here X_s=synchronizing reactance of each machine

$\therefore$ synchronizing power supplied by machine 1 is $P_{sy} = E_1 I_{sy} = E \times \frac{\alpha E}{2X_s} = \frac{\alpha E^2}{2X_s}$

$$P_{sy} = \frac{\alpha E^2}{2X_s}$$

Total synchronizing power for three phases $= 3\,P_{sy} = \frac{3\alpha E^2}{2X_s}$

Note that this is the value of synchronizing power when two alternators operate in parallel at no-load.

……………………………..xxxxxxxxxxxxxxxxxxxxxxxxx……………………………

CHAPTER-III

SYNCHRONOUS MOTOR

Defination :

A synchronous Motor is electrically identical with an alternation.

Characteristic Feature :→

(1) It runs either at synchronous speed or not at all. The only way to change its speed is to very the supply frequency $\left(\because Ns = \frac{120f}{P}\right)$.

(2) It is not inherently self starting. It has be run upto synchronous (or rearly synchronous) speed by means before it can be synchronorised to the supply

(3) It is capable of being operated under a wide range of power factor both lagging and leadig.

Construction :→

Like an alternator, a synchronous motor has the following main parts :

(i) A stator which houses 3-phase distributed armature winding in the slots of the stator core and receives powers from a 3ϕ -supply.

(ii) A rotor that has a set of salient poles excited by direct current to form alternator N and S poles.

The exciting coil are connected in series to the slip rings and direct current is fed in to the winding from an external excite on mounted on the rotor shaft.

The stator is wound for the same number of poles as the rotor poles

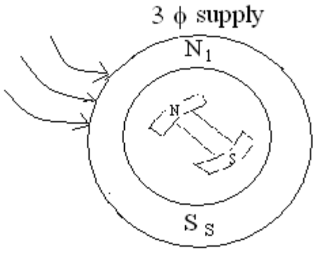


Fig (a)

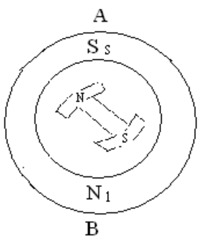


Fig - (b)

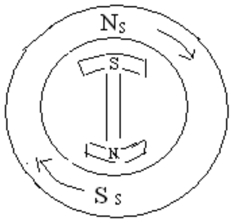


Fig - (c)

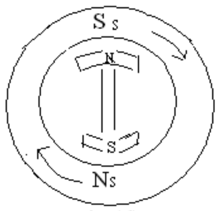


Fig-(d)

Principle of Operation :→

When a 3ϕ winding is fed by a 3ϕ supply then a magnetic field of constant magnitude but rotating synchronous speed is produced in the stator. Consider a two pole stator which are shown two stator pole Ns and Ss rotating at synchronous speed in clockwise direction.

With the rotor position as shown in fig. suppose the stator pole are at that instant situated at point B. The two similar poles N and Ns as well as S and Ss will repel each other with the result that the rotor tends to rotate on the anticlockwise direction.

But half a period later stator poles having rotate around interchange their position i.e. Ns is at point B and Ss is at point A. Under these condition Ns atleast S and Ss attract N. Hence rotor tends to rotate clockwise (which is just the reverse direction).

It is seen that due to continuous and rapid rotation the stator pole, te rotor is subjected a torque which is rapidly revenging owing to this large inertial, the rotor cannot instantaneously respond to such likely revenging torque with the result that it remains stationary.

Considering the condition in fig(b) the stator and rotor are alternating each other. Suppose that the rotor is not stationary but is rotating clockwise with such a speed that it rotates through one pole pitch by the times the stator poles interchanged their positions. As shown in fig(c). Hence again stator and rotor poles attract each other. It means that if the rotor pole also shift their position along with the stator poles, then they will produce a torque.

Method of starting :

The following steps are adopted for starting of a synchronous motor.

(1) The fed winding is shorted that means D.C. excitation is not given to the field winding.

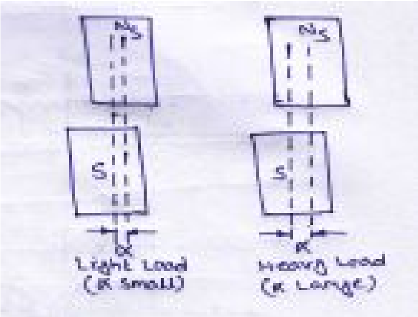
(2) Reduced voltage with the help of auto-trans transformer is applied across the stator terminal.

(3) When the motors at far nearly 9% of the synchronous speed, a d.c. excitation as applied to the field winding. At that time the motor is synchronized.

(4) full supply voltage is applied across stator terminals by cutting out the autotransformer .

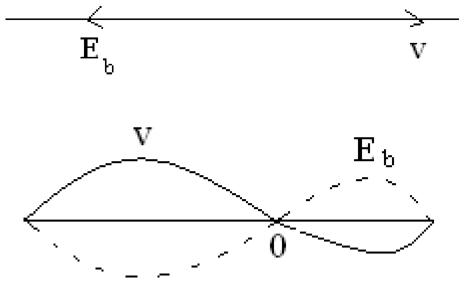
(5) the motor then can be operated at any power factor by the d.c. excitation.

(6) At light load or heavy load condition the rotor advances and backs to the stator flux respectively with an angle α. It is called the load angle.



Motor on no-load :

When a 3ϕ supply is feed to a synchronous motor, the motor starts rotating. As a result back emf (E_s) is set up in the alternator (stator) by the rotor flux which oppose the applied voltage 'V' his back emf of (Es) depends on rotor excitation only (and not on peed) as in D.C. motor. The net voltage of armature (stator in the voter difference of V & E_b Armature current is obtained by dividing this vector difference of voltage by the armature impendance

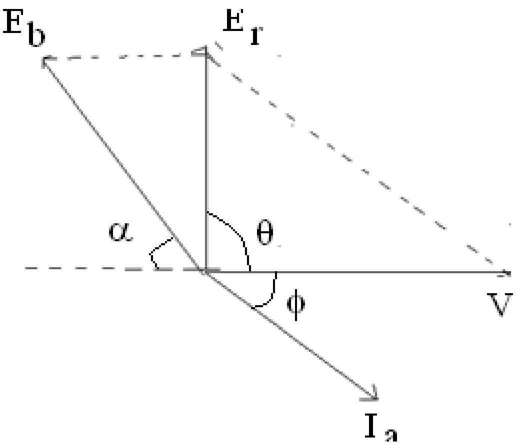


i.e. $I_a = \dfrac{\vec{V} - \vec{E_b}}{S}$

The Fig, show condition when the moor is running on no- direction on no-load & is having loss. As field excitation is given such that Eb = V

The vector difference of E_b & V is f hence hence the armature current is also zero)

If the motor is on no load but it has losses then θ vector for Eb faces back by a certain angle 'α' . So that a resultant voltage E_R & hence current 'I_a is brought into existence which supplies the losses.

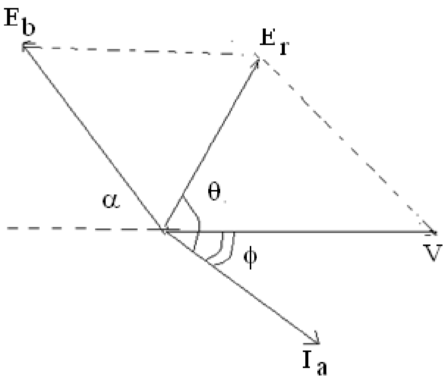


Motor on Load :

(Effect of varying load with constant excitation)

If the load motor is landed then the back emf (E_b) places back by a certain value called the "load angle" or coupling angle.

The net voltage current the armature $E_R = \vec{V} - E_b$

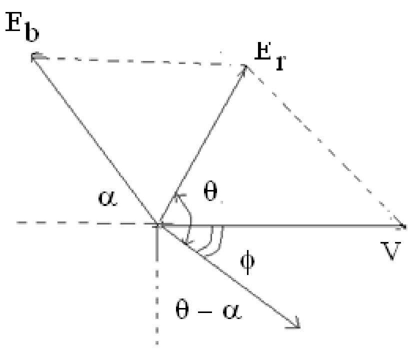


If the motor is further loaded, its rotor further back by a greater value of load angle 'α' the resultant voltage 'E_R' is increased and the motor draws an increased armature current but at a devised power factor.

$$I_a = \frac{\vec{V} - \vec{E}_b}{Z_S} = \frac{\vec{V} - \vec{E}_b}{R_a + jX_s}$$

$$\Rightarrow V = E_b + I_a Z_s$$

$$= E_b + I_a(R + jX_S$$



Where Z_S = Synchronous impedance / phase

R_a = Armature resistance / phase

X_S = Synchronous reactance / phase

The angle ‘θ’ is known as “ initial angle” by which ‘I_a’ lags behind E_R.

$$\theta = \tan^{-1}\frac{X_s}{R_a}$$

ϕ is the phase angle by which ‘I_a’ lags behind ‘V’.

If R_a is negligible , then θ = 90°

Motor I/p per phase = $VI_a \cos\phi$

Total I/p to the motor = $\sqrt{3}V_L I_L \cos\phi$

Mechanical power developed in the rotor per phase

$$P_m = E_b I_a \cos(\alpha - \phi)$$

Different Power stages in a Synchronous Motor

Power I/p per phase

[$VI_a \cos\phi$]

Armature cu loss

(I_{a2} Ra) and iron loss

Mechanical power in

armature Pm = $E_b I_a \cos(\alpha - \theta)$

Iron loss and friction losses

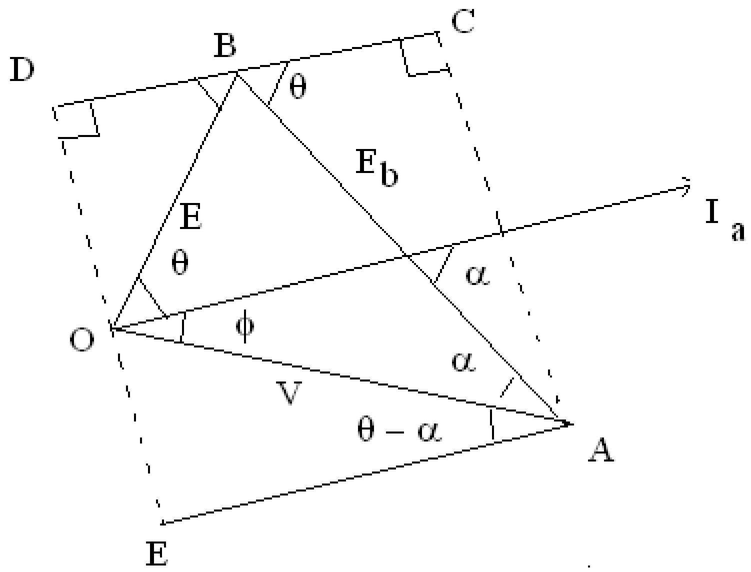
O/p Power in B.H.P.

Power or Torque Developed by the Motor :

OA represents supply voltage/phase. The armature current per phase = I_a AB is the back emf at a load angle of α OB gives the resultant voltage $E_R = IZ_s$. 'I' leads 'V' by angle 'ϕ' and lags behind 'E_R' by a angle 'θ'.

$$\theta = \tan \frac{x_3}{R_a}$$

Line CD is drawn at an angle θ to AB, AC and ED are perpendicular to CD (and hence to AE also)



Mechanical power per phase developed in the rotor is

$P_m = E_b I \cos\Psi$ ---------------- (i)

Fn ΔOBD, BD = $IZ_s \cos \Psi$

BD = CD-BC = AE-BC

IZs cosΨ = V cos (θ-α)-E_b cos

$$\Rightarrow I \cos \psi = \frac{V}{Z_s} \cos (\theta - \alpha) - \frac{E_b}{Z_s} \cos\theta$$

Substituting this value in equation (1), we get Pm per phase.

$$P_m = E_b \left[\frac{V}{Z_s} \cos(\theta - \alpha) - \frac{E_b}{Z_s} \cos\theta) \right]$$

i.e. $$\boxed{P_m = \frac{E_b V}{Z_s} \cos(\theta - \alpha) - \frac{E_b{}^2}{Z_s} \cos\theta}$$

Maximum power developed depends on the load angle 'α'. So condition for maximum power developed can be found by differential long P_m wants α and then equating it to zero

$$\frac{dP_m}{d\alpha} = 0$$

$$\Rightarrow \frac{d}{d\alpha}\left[\frac{E_b V}{Z_s}\cos(\theta-\alpha) - \frac{E_b^{\ 2}}{Z_s}\cos\theta\right] = 0$$

$$\Rightarrow \frac{E_b V}{Z_s}[-\sin(\theta-\alpha)) - 0 = 0$$

$$\Rightarrow \frac{E_b V}{Z_s} - \sin(\theta-\alpha) = 0$$

$\Rightarrow \sin(\theta\text{-}\alpha) = 0$

$\Rightarrow \sin(\theta\text{-}\alpha) = \sin 0$

$\Rightarrow \theta - \alpha = 0$

$\Rightarrow \theta = \alpha$

Value of maximum power

$$(P_m)_{\max} = \frac{V_{Eb}}{Z_s}\cos(\theta-\alpha) - \frac{E_b^{\ 2}}{Z_s}\cos\alpha$$

$$= \frac{V_{Eb}}{Z_s} - \frac{E_b^{\ 2}}{Z_s}\cos\alpha$$

$$= \frac{V_{Eb}}{Z_s} - \frac{E_b^{\ 2}}{Z_s}\cos\theta$$

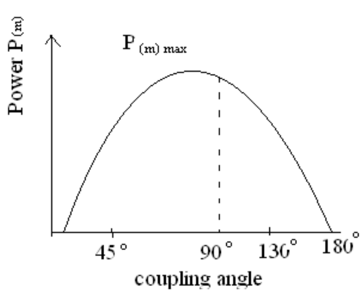
Now, maximum power developed in the rotor depends on the value of internal angle ‘θ’.

$\rightarrow$ When $\theta = \alpha = 90°$

The power will be maximum.

$$(P_m)_{\max} = \frac{VE_b}{Z_s}$$

If ‘T_g’ is the grass rotor torque developed by the motor, then



$$\left.\begin{aligned} T_g \times 2\pi N_s &= P_m \\ \Rightarrow \quad T_g &= \frac{P_m}{2\pi N_s} \end{aligned}\right\} N_s \ in \ rps$$

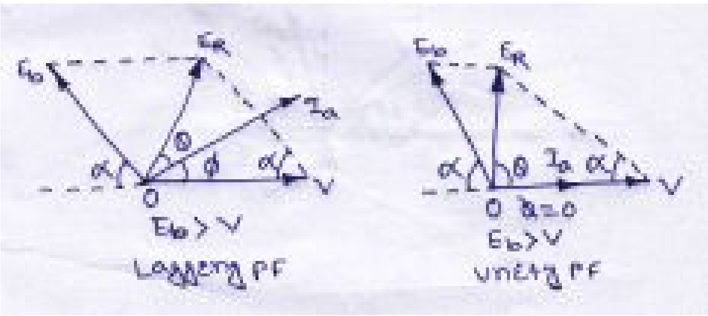
$$\Rightarrow \quad T_g = \frac{P_m}{2\pi N_s / 60}[N_s \ in \ rpm]$$

$$\Rightarrow \quad T_g = \frac{60}{2\pi} \times \frac{P_m}{N_s} = \boxed{9.55 \frac{P_m}{N_s} N-m}$$

Shaft torque is given as

$$\boxed{T_{sh} = 9.55 \frac{P_{oul}}{N_s}}$$ N-M

Synchronous Motor with different excitations



(i) When $E_b = V$, is known as normal excitation

(ii) When $E_b < V$, is known as under excitation, lagging pf .

(iii) When $E_b > V$, is known as over excitation. leading pf.

The diagram shows the induced back emf in various pf.

i) **Lagging pf:**

$$AC^2=AB^2+BC^2=[V-E_R\cos(\theta-\varphi)]^2+[E_R\sin(\theta-\varphi)]^2$$

$$E_b=\sqrt{[V - IaZs\cos(\theta - \varphi)]^2 + [IaZs\sin(\theta - \varphi)]2}$$

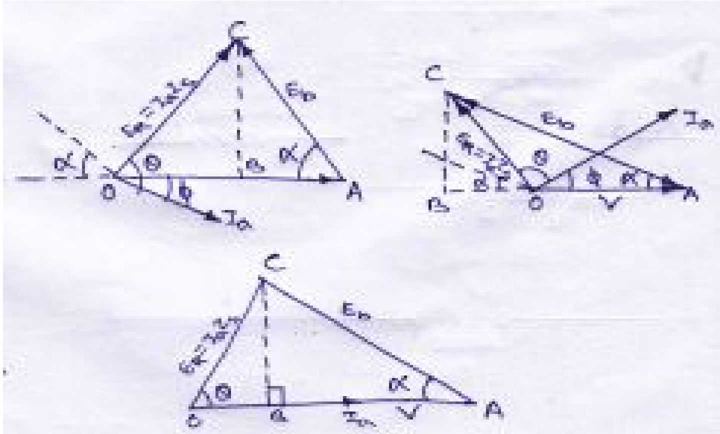
ii) **Leading pf:**

$$E_b = V + I_aZ_s\cos[180^0 - (\theta + \varphi)] + jI_aZ_s\sin[180^0 - (\theta + \varphi)]$$

iii) **Unity pf:**

$OB=I_aR_a$, $BC=I_aX_s$

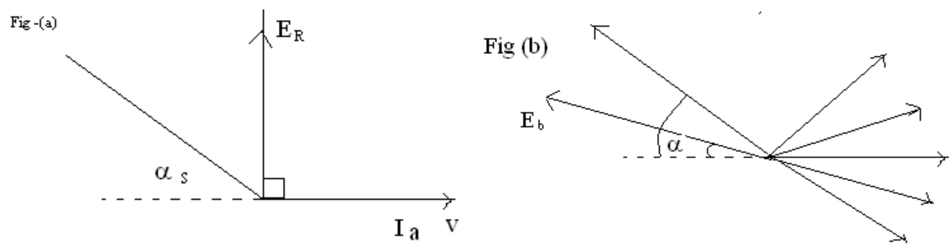
$E_b=(V-I_aR_a)+jI_aX_s$

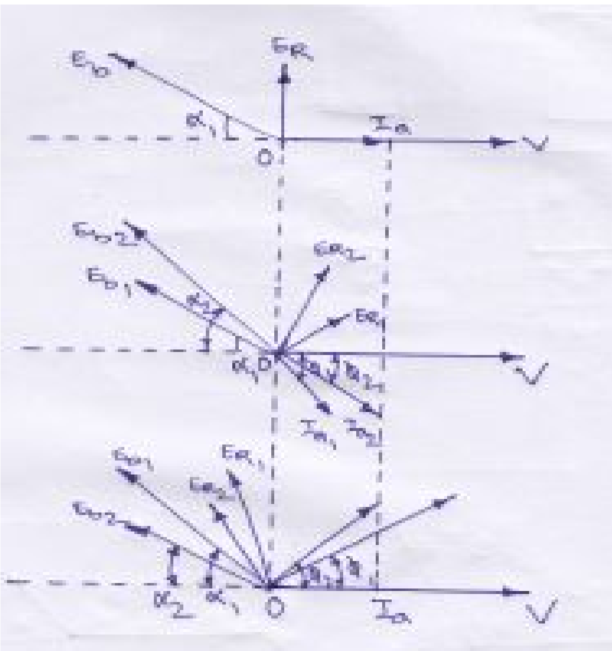


Motor on Load

⇒ Effect of varying excitation with constant load.

Suppose a synchronous motor is operating with normal excitation (E_b=V) at unity power factor with a given load. The armature is drowning a power of Via per phase which is enough to meet the mechanical load on the motor. The effect of changing excitation with load remains constant is discussed below.





Excitation Decreases:

As shown in fig (1) suppose due to decrease in excitation, back emf is reduced to E_{n_1}, at the same load angle 'α'. The resultant voltage E_{R_1}, causes a lagging armature current I_{a_1},to flow. Even though I_{a_1} is larger than I_{a_2} in magnitude. It is capable of producing necessary VI_a for carrying the constant load because $I_{a_1} \cos\phi$. Component is less than I_a so that $VI_a \cos\phi_1 < VI_a$.

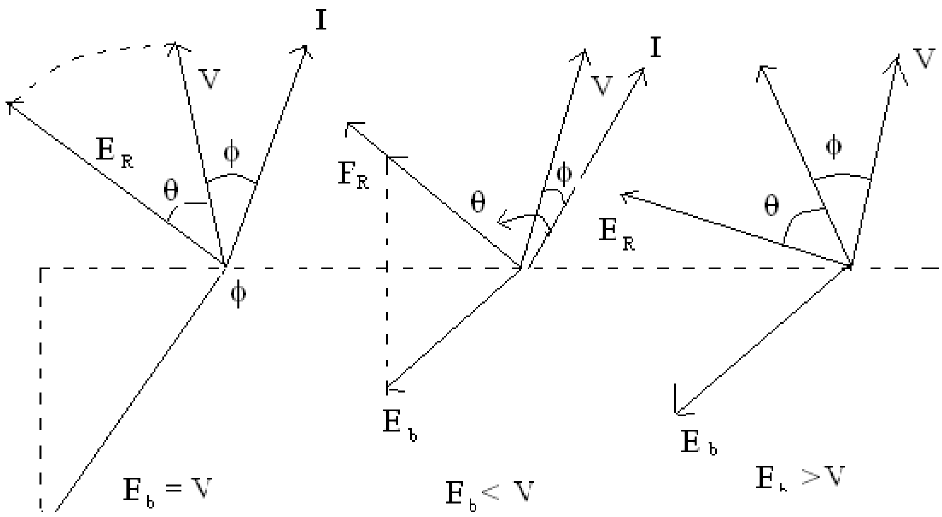
Hence, it becomes necessary for load angle to increase from α_1 to α_2, it increase back emf from E_{b_1} to E_{b_2},which increases resultant voltage from E_{R_1} to E_{R_2} consequently, the armature current increases to I_{a_2} whose in phage component.

(b) Excitation Increased

The effect of increasing field excitation is shown in fig (c), where increased E_n is shown at the original load angle α_1 The resultant voltage causes loading current I_{a_1} whose in phase component is larger than I_a. Hence armature develops more power than the load in the motor. Accordingly, load angle decreases from α_1 to α_2 which develops resultant voltage from E_{R_1} to E_{R_2} consequently armature current decreases from I_{a_1} to I_{a_2} whose in phase component. $I_{a_2} \cos\phi_2 = I_a$. In the case armature current develops power sufficient to carry the constant load on the motor.

Hence, it is seen that variations in the excitation of the synchronous motor running with given load. Produces variation in its load angle only.

Effect of Excitation on Armature Current and Power Factor.



Consider a synchronous motor in which mechanical load is constant.

Fig (a) shows the case for 100% excitation i.e. when $E_b = V$

The armature current I lags behind 'V' by an small angle φ with E_R is fixed by stator constants i.e. $\tan\theta = \frac{x_s}{R_a}$. In fig (b), the excitation is loss then 100% $E_b < V$. Here E_R is advanced in clock wise direction & also armature current. The magnitude of 'I' is increased but its power factor is decreased. The component of Icosφ remains the same as before but wattless component I sin φ is increased.

As excitation is decreased I will increased but p.f. will decreased.

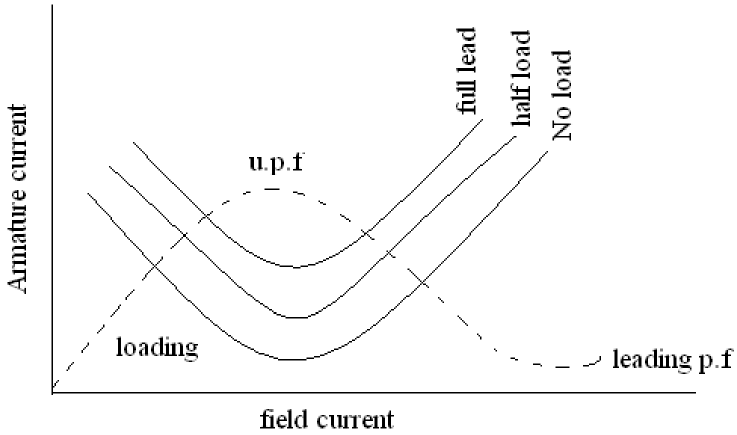
Fig. (c) represents the conduction for over excited motor i.e. $E_b > V$. Here the resultant voltage E_R is pulled in anticlockwise direction & also 'I'. Now the motor is drawing a loading current.

for same voltage of excitation, 'I' will be in phase with 'v' i.e. p.f. is unity AC that the current drawn by the motor would be maximum.

V-Curve.

Magnitude of armature current varies with excitation. The current has large value for low & high value of excitation. In between them, at any certain excitation. The current is minimum. The variation of current 'I' with excitation is shown in fig below which is known as V-curve. Since it looks like 'V'.

Motor runs with loading p.f. when over-excited and with lagging p.f. when under excited. In between, the power factor will be unity. The variation of p.f. with excitation (shown in dotted line) which is known as inverted v-curve, since is looks like inverted 'V'. It is seen that the armature current will be minimum when the p.f. will be unity.



Synchronous Condenser / Capacitor

Both transformer & induction motor draws lagging currents from the lien on light loads, the power drawn by them has a large reactive component and power factor is very low.

It is seen that an over-excited synchronous motor will run with leading p.f. by using synchronous motor and transformers. The power factor can be increased and the reactive components of power can be decreased when synchronous motors are used for this purpose (p.f. improvement), is known as synchronous condenser or synchronous capacitor because it draws a leading current from the line like a capacitors because it draws a leading current from the line like a capacitor.

Hunting or Phase Swinging.

When synchronous motor is used for driving a varying load. Then a condition known as hunting is produced. Hunting may also be caused if supply frequency is pulsating.

When a synchronous motor is loaded its rotor falls back in phase by the coupling angle 'α'. As the load is progressively increased, thus angle also increases so as to produced motor torque to meet with the increased load. If now there is a sudden decrease in the motor load, the rotor is pull back to new value of 'α'. In this way rotor starts oscillating about its new position of equilibrium. Corresponding to the new load. If the time period of those oscillations equals to the natural time period of the machine, then mechanical resonance is set up. The amplitude of these oscillations may become so large to throw the machine out of synchronous.

To stop the oscillations damper or damping grids are employed. These damper-consists of (short circuited copper were in the faces of the field poles of the motor. The oscillatory motion of the rotor sets up eddy currents in the dampers which flows in such a way as to suppress the oscillations.

Comparison between induction motor and synchronous motor.

Induction Motor		Synchronous Motor	
1.	Its speed decreases slightly as the load increased.	1.	It runs at either synchronous speed or not at all
2.	It always run with lagging power factor.	2.	It can be run under a wide rang of p.f. both lagging & leading.
3.	It is self –starting.	3.	It is inherently not self-starting.
4.	D.C. excitation is not required	4.	D.C. excitation is required.
5.	It is cheap & simple	5.	It is costlier & complicated.

Application of Synchronous Motor :

1. Power factor correction :

Over excited synchronous motor having leading p.f. are widely use for improving p.f. of power systems.

2. Constant speed application.

High speed synchronous motor (above 600rpm) are used for centrifugal pumps, belt driven reciprocating, compressor, blowers, line shafts, rubber & paper mills etc.

Low speed synchronous motors (below 600rpm) are used for drives such as centrifugal and screw type pumps, balls and tube mills, vaccum pumps, choppers and metal rolling mills etc.

3. Voltage regulation :

The voltage at the end of a long transmission line varies greatly when large inductive loads are present. By installing a synchronous motor with a field regulator, this voltage rise can be controlled.

By varying the excitation, the p.f. can be made lagging on leading which helps to maintain the lien voltage at its normal value.

Example :- 1

A 75-kW, 3-ϕ, Y-connected, 50-Hz. 440-V cylindrical rotor synchronous motor operates at rated condition with 0.8 p.f. leading. The motor efficiency excluding field and stator losses, is 95% and X_5 = 2.5Ω. Calculate (i) mechanical power developed (ii) armature current (iii) back e.m.f. (iv) power angle and (v0 maximum or pull-out torque of the motor.

Solution : $N_S = 120 \times 50/4 = 1500$ rpm = 25 rps

(i) $P_m = P_{in} = P_{out} / \eta = 75 \times 10^3/0.95 = 78.950$ W

(ii) Since power input is known

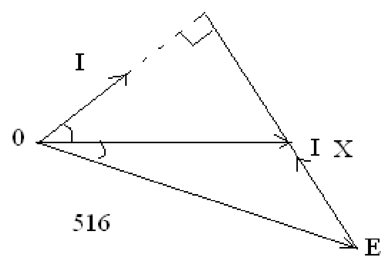
$\therefore \sqrt{3} \times 440 \times I_a \times 0.8 = 78.950;$ $I_a = 129$ A

(iii) Applied voltage /phase = $440/\sqrt{3}$ = 254 V. Let V = 254 $\angle 0°$ as shown in fig

Now, $V = E_b + jIX_S$ or $E_b = V - jI_aX_S = 254 \angle 0° - 129$

$\angle 36.9° \times 2.5 \angle 90° = 250 \angle 0° - 322 \angle 126.9° = 254 - 322$

$(\cos 126.9° + \sin 126.9°) = 254 - 322 (-0.6 + j\,0.8) = 516 \angle -30°$



iv) $\therefore$ $\alpha = -30°$

v) pull –out torque occurs when = $\alpha = 90°$

maximum $P_m = 3\frac{E_b V}{X_S}\sin\delta = 3\frac{256\times 516}{2.5} = \sin 90^o = 157{,}275W$

$\therefore$ pull-out torque = 9.55 × 157, 275/1500 = 1,000 N-m.

Example : 2

A 20-pole, 693-V, 50Hz, 3ϕ, Δ- connected synchronous motor is operating at no-load with normal excitation. It has armature ressistance per phase of zero and synchronous reactance of 10Ω. If rotor is retarded by 0.5°(mechanical) from its synchronous position, compute.

(i) rotor displacement in electrical degrees

(ii) armature emf / phase

(iii) armature current / phase

(iv) power drawn by the motor

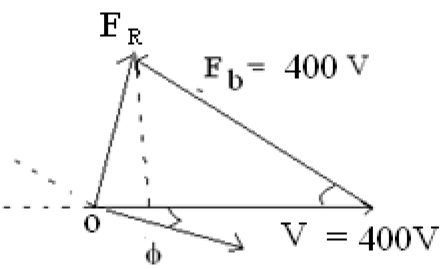
(v) power developed by armature

How will these quantities change when motor is loaded and the rotor displacement increases to 5° (mechanical) ?

Solution :

(i) $\alpha(elect.) = \frac{P}{2}\times\alpha(mech)$

$\therefore \alpha(elect) = \frac{20}{2}\times 0.5 = 5^o(elect)$



(ii) $V_P = V_L/\sqrt{3} = 693/\sqrt{3}$

$= 400V,$

$\therefore E_R = V_p - E_b\cos\alpha) + jE_b\sin\alpha = (400 - 400\cos 5^o + j400\sin 5^o)$

= 1.5 + j 35 = 35 ∠ 87.5 V / phase

(iii) $Z_S = 0 + j10 = 10\angle 90°$; $I_a = E_R / Z_S = 35\angle 87.5°/10\angle 90°$

$= 3.5\angle -2.5°$ A/ phase

Obviously, I_a lags behind V_p by 2.5 °

(iv) Power input/ phase $V_p\, I_a \cos\phi = 400\times 3.5\times\cos 2.5° = 1399$ W

Total input power = $3\times 1399 = 4197$ W

(v) Since Ra is negligible, armature Cu loss is also negligible. Hence 4197 W also represent power developed by armature.

Example – 3

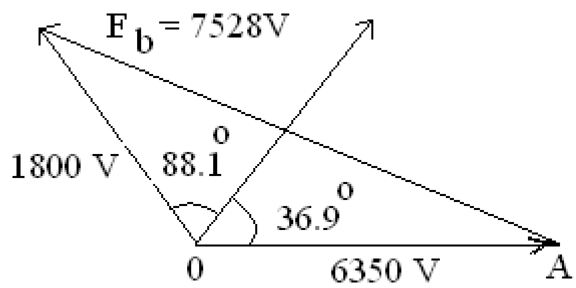
The input to an 11000-V, 3-phase, star-connected synchronous motor is 60 A. The effective resistance and synchronous reactance per phase are respectively 1 ohm and 30 ohm. Find (i) the power supplied to the motor (ii) mechanical power developed and (iii) induced emf for a power factor of 0.8 leading.

Solution : (i) Motor power input = $\sqrt{3}\times 11000\times 60\times 0.8 = 915kW$

(i) star Cu loss/phase = $60^2\times 1 = 3600W$; Cu loss for three phase = $3\times 3600 = 10.8kW$

$P_m = P_2 -$ rotor Cu loss = 915 – 10.8 = 904.2 kW

$V_P = 11000\sqrt{3} = 6350V, \phi = \cos^{-1}0.8 = 36.9^o;$



$\theta = \tan^{-1}(30/1) = 88.1^o;$

$Z_S = 30\Omega;$ stator impedance drop/ phase = $I_a Z_s$

= 60 × 30 = 1800 V

As seen from Fig. 38.25

$$E_b{}^2 = 6350^2 + 1800^2 - 2\times 6350\times 1800\times\cos(88.1^o + 36.9^o)$$

$= 6350^2 + 1800^2 - 2\times 6350\times 1800\times -0.572$

$\therefore E_b =$ 7528 V; line value of E_b = 7528 ×√3 = 1.3042.

Example -4 : A 500-V, 1-phase synchronous motor gives a net output mechanical power of 7.46 kW and operates at 0.9 p.f. lagging. Its effective resistance is 0.8 Ω. If the iron and friction losses are 500 W and excitation losses are 800 W, estimate the armature current. Calculate the commercial efficiency.

Solution : Motor input = $VI_a \cos\phi$; Armature Cu loss = $I_aR_a{}^2$ Power developed in armature is $P_m = VI_a\cos\phi - I_aR_a{}^2$

$\therefore I_a{}^2R_a - VI_a\cos\phi + P_m = 0$ or $\therefore I_a = \dfrac{V\cos\phi \pm\sqrt{V^2\cos^2\phi - 4R_aP_m}}{2R_a}$

Now, P_{out} = 7.46 kW = 7,460 W

$P_m = P_{out}$ + iron and friction losses + excitation losses

= 7460 + 500 + 800 = 8760 W

$$I_a = \frac{500\times 0.9 \pm \sqrt{(500\times 0.9)^2 - 4\times 0.8\times 3760}}{2\times 0.8} = \frac{450\pm 417.7}{1.6} = \frac{32.3}{1.6} = 20.2A$$

Example :- 5

The synchronous reactance per phase of a 3-phase star-connected 6.600 V synchronous motor is 10Ω. For a certain load , the input is 900 kW and the induced line emf is 8,900 V. (line value) Evaluate the line current. Neglect resistance.

Solution : Applied voltage / phase = 6.600/ √3 = 3, 810 V

Back e.m.f. / phase = 8,900 / √ 3 = 5, 140 V

Input = $\sqrt{3}V_L.I\cos\phi = 900.000$

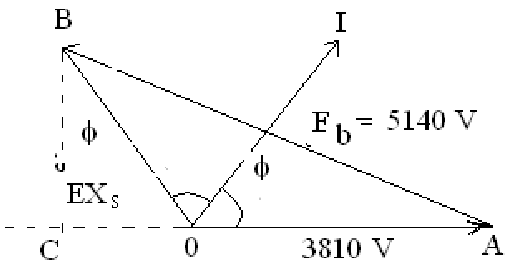
∴ $I\cos\phi$ = 9× 105 / √3×6.600 = 78.74 A

In Δ ABC of vector digram in Fig we have $AB^2 = AC^2 + BC^2$

Now OB = $I.X_S$ = 10 I

BC = OB cos ϕ = 10 × 78.74 = 787.4 V

∴ $5{,}140^2 = 787.42 + AC^2$



∴ AC = 5,079 V

OC = 5,079 – 3, 810 = 1, 269 V

Tan ϕ = 1269/787.4 = 1.612 ϕ = 58.2°, cosϕ = 0527

Now I cos ϕ = 78.74; I = 78.74/0.527 = 149.4 A

CHAPTER-IV

Single Phase induction Motor

A single phase Induction Motor (I.M.) is very similar to 3 Phase squirrel cage I.M. It has a squirrel cage rotor and a single phase winding on stator like 3 phase I.M., single phase I.M. is not self starting. The stator winding produces a magnetic field which polarity reversed after each half cycle. So the field don't produce rotating field. If a single phase I.M. having squirrel cage rotor and 1-phase distributed stator winding, it doesn't develop any resulting starting torque as the torque developed in both the cycle neutralize each other. To make the I.M. starting, we have to add an another winding in the stator circuit is known as auxiliary winding (starting)

Making Single Phase I.M. Self starting :

To make a 1-phase I.M. self starting we should some how produce a revolving stator magnetic field, this may be achieved by converting a 1-phase supply in to two phase supply by using an additional winding. Hence the rotor of the single phase motor starts rotating like 3 phase motor. When it achieves sufficient speed, the additional winding may be removed. But the rotor continue running.

Different types of single phase I.M.-

1. Induction Motors like split-phase, capacitor and shaded pole type.
2. Repulsion type motors
3. A.C. series motors (Commutator motors) etc.

Split Phase Motor:

The Stator circuit of a split phase I.M. is added with an auxiliary winding with the main winding and it is located 90^0 electrically apart from the main winding. The two windings are so designed that the auxiliary winding has high resistance and small reactance while the main winding has low resistance and large reactance.

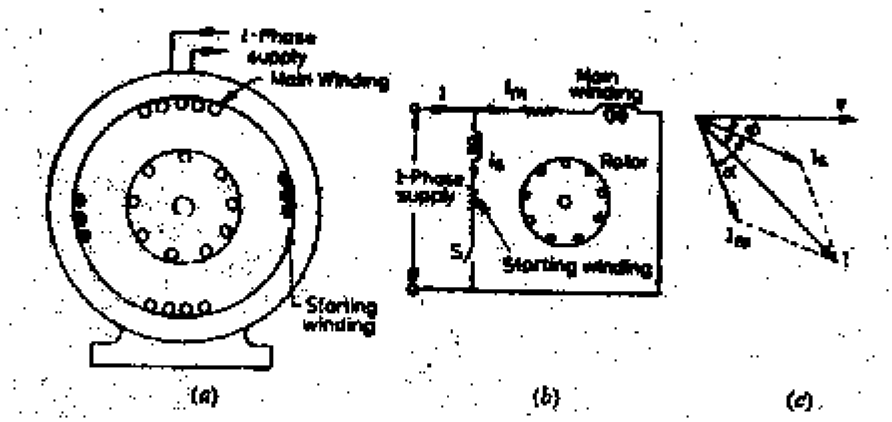


Figure 1 split phase I M motor

Operation-

When supply is given to the starter windings both the windings are energized. Since main winding is made by highly inductive while the auxiliary winding is resistive that produce a weak revolving field for which it produces revolving flux and rotor starts revolving hence the motor started.

The Torque produced is,

$$Ts = K\, I_m\, I_s \,\text{Sin}\, \alpha$$

When α is the phase angle between I_m & I_s. When the motor achieves about 75 % of synchronous speed, the centrifugal switch S will open and the auxiliary winding is cut off from the circuit. Then the motor operates as a $1 - \Phi$ I.M. and it continues to accelerate till it reaches it's normal speed which is below the synchronous speed. The starting torque is proportional to the Current

If the starting period delay exceeds 5 Seconds, the winding may burn out because the winding made of fine wire.

Uses

Fan, Washing machine, small machine tools etc.

Capacitor Start I.M. :

A Capacitor start motor is identical to a split phase motor except that the starting winding has same number of turns as main winding and a capacitor is connected in series with the starting winding.

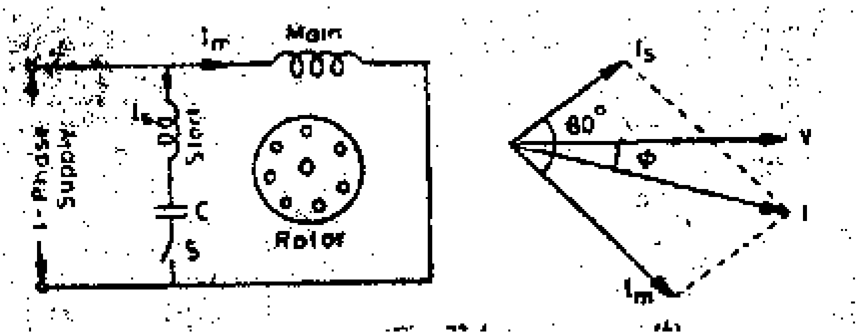


Figure 2 Capacitor Start I.M.

Operation

The value of the capacitor is such that “I_s” leads “I_m” by The starting torque which is more than the split phase I.M. When torque is produced, the rotor starts rotating. When the rotor achieves 75 % of the N_s, the centrifugal switch will be open. Then auxiliary winding is cut off from the circuit. The motor then operate as a 1-phase I.M. and continue to accelerate till it reaches it’s normal speed.

Advantages

It’s starting characteristics are better than the split phase I.M. For the same starting torque, the current of starting winding is only about half that in split phase I.M. so, it is heated less quickly.

Uses :

It is used where low starting torque is required.

Capacitor start and run

It is similar to capacitor start motor except that the starting winding is not opened after starting. So, when the motor runs both windings are connected in the circuit . It has two capacitors with the starting winding. The capacitor C_1 has smaller capacity than C_2 and is connected in the circuit in series with the starting winding permanently during starting as well as running. The large capacitor C_2 is connected in parallel in C_1 for starting purpose only. When the motor approaches about 75 % of N_s then Centrifugal switch is opened and the capacitor C_2 is disconnected from circuit.

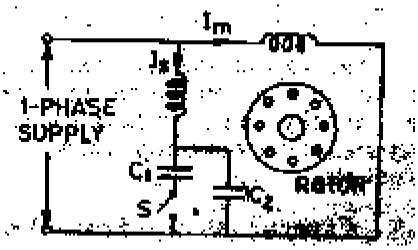


Figure 3 capacitor start induction motor

The important kind of capacitor motor is permanent capacitor motor. In such type the capacitor is permanently connected to the circuit and one in number only.

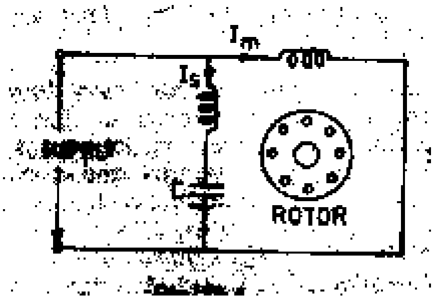


Figure 4 permanent capacitor motor

Characteristics

This type of motor is designed for perfect 2–phase operation at any load and it produces continuous torque as compared to induction motor.

Uses

Due to it's continuous torque and vibration free, it is used in hospitals, studio, refrigerators, compressors, stokers, ceiling fan, blowers etc.

Shaded Pole Motor

The shaded pole motor is very popular for rating up to 0.05 HP. A small portion of pole core of about 30% is slot cut and surrounded by a short circuited ring of Cu strip called shading coil.

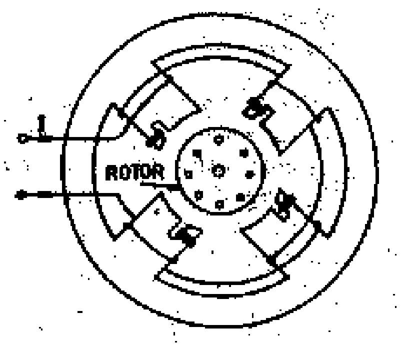


Figure 5 shaded pole motor

Operation

From the total of core, the flux produced and emf is induced in the shading coil. The resulting current in shading coil is in such a direction, so as to oppose I and so the charge in flux according to Lens's law. So this flux in the shaded portion of the pole is weakened while in the unshaded portion is strengthened. The magnetic axis lies along the middle of this part.

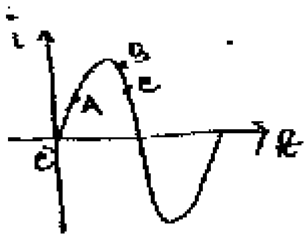


Figure 6 torque in shaded pole motor

During the portion in AB as shown in figure (6), the flux is reached almost maximum value, the flux distribution across the pole is uniform. Since no current is flowing in shading coil, the magnetic axis shift to the centre of the pole.

As the flux decreases as shown in figure (6), from B to C, This again set a induced current in the shading coil. This current flows in such a direction that to oppose the decrease in current. Thus the flux in the shaded portion of the pole is strengthened while the unshaded portion is weakened. So the magnetic axis shift to the middle part of the shaded pole.

This shifting of flux is like a rotating weak field moving in the direction from unshaded portion to shaded portion of the pole. Under the influence of the moving field a small staring torque is developed which torque starts to rotate the rotor, additional torque is produced by single phase motor action. Such motors are built in very small sizes of 5-50w but are simple in construction and are extremely rugged, reliable and cheap.they do not need any commutator, switch, brush, collector rings etc. However they suffer from disadvantages of (i) low starting torque, (ii) very little over load capacity and (iii) very low efficiency ranging from 5% to 35% from lower to higher ratings respectively.

Uses

It is used in small fans, toys, hair drier of power up to 50 W.

AC Series Motor / Universal Motor

The construction of AC series motor is as like as DC series motors. If a DC Series motor is connected to an AC supply, it will rotate and produce unidirectional torque because the current flowing in both the armature and field reverses at the same time. When a DC series motor operates on a single phase supply, then it is called a AC series motor. The performances of this type of motor will not be satisfactory due to the following reasons.

1. The alternating flux would cause excessive eddy current loss in the yoke and the field core will become extremely heated.
2. Sparking will occur at brushes because of huge voltage and current induced in the short circuited armature coil during commutation period.
3. Power factor is very low.

   Due to the above drawbacks DC series motor required some changes by which AC supply input disadvantages solved.. The changes made are

a) The entire magnetic circuit is laminated in order to reduce the eddy current losses.
b) A high field flux is obtained by using a low reluctance magnetic circuit.
c) Excessive sparking eliminated by using high resistance leads to connect the coil to the commutator segment.

Though this type of motor can be operated either on AC or DC supply, the resulting torque speed curve is same. It is also known as Universal motor.

Operation

When it is connected to an AC supply the same alternating current flows through the field and armature winding. The field winding produces an alternating flux that react with an armature current to produce a torque and the direction of the torque is always same because they (current and flux) reverses simultaneously.

Characteristics

a) Speed increases to a high value with a decreasing load.
b) It has very high starting torque.
c) At Full load, the power factor is 90 %.

   Uses.

   a. Sewing machine b. vacuum cleaners, c. mixer grinders and blenders

d) High speed vacuum cleaners. e. hair driers f. power saw

f. Drills g. Electric Shaver.

Single Phase Repulsion Motor

A repulsion motor is similar to an Ac Series motor except some modification. The brushes are not connected to supply but are short circuited by themselves. The current induced in the armature conductor by mutual induction method.

Construction **Single Phase Repulsion Motor**

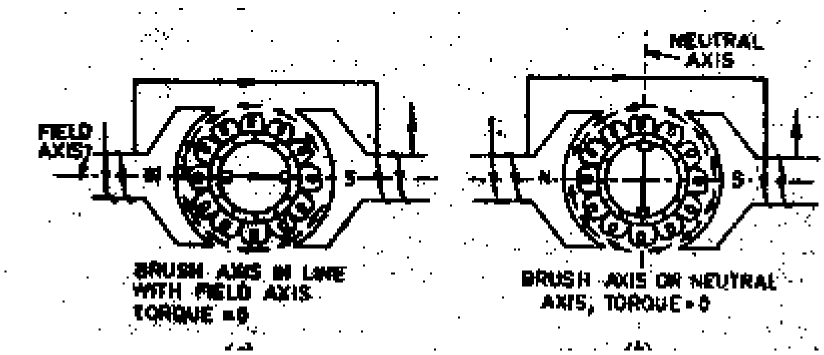


Figure 7 Single Phase Repulsion Motor

The field of the stator winding is connected directly to the AC single phase source. The rotor is similar to a DC motor armature winding connected to the commutator. The brushes are short circuited which make the rotor squirrel cage type. It has very high starting torque and also better power factor as compared to other single phase motor.

Operation

The figure shows, two pole repulsion motor with short circuited brushes. The brush axis is parallel to stator field. The emf is induced in the armature Conductor by induction method and current flows through the rotor conductors. The current flows from N to S brush in two paths. during this brush position half of the rotor conductors under N pole carry current inward and half carry current outward. The same thing occurs under S pole. Therefore, same torque is produced in opposite direction in both the half coils. So the net torque is zero.

If the brush axis is in some angle other than 0^0 or $90^{0,}$ then a torque is developed in the rotor and accelerate the rotor to final speed. The brush axis is shifted in clockwise direction through some angle from stator field axis. The emf is induced in same direction, the current flows in two paths of the rotor winding between N & S. Now the more conductors under North pole carrying current in one direction while more conductors under south pole carrying current in opposite direction, so that the torque is developed in clockwise direction and the rotor rotates to it's final speed.

The direction of rotation of the rotor depends upon the direction in which the brushes are shifted. If the brushes are shifted in clockwise direction from the stator field axis then the net torque in clockwise direction. It has high starting torque.

Use

Commercial refrigerators, compressors and pumps.

CHAPTER -V

COMMUTATOR MOTORS

A.C. Series Motor or Universal Motor

A dc series motor will rotate in the same direction regardless of the polarity of the supply.

When a dc series motor operates on a single phase ac supply it is called an AC series motor. However some changes are required in a DC motor so that it can satisfactorily operates on A.C. supply.

The changes are as follows:

i) The field core is constructed of a material having low hysteresis loss. It is laminated in order to reduce eddy current loss. Hence A.C. series motor requires a more expensive construction than a D.C. series motor.

ii) The series field winding uses as few turns as possible to reduce the reactance of the field winding to minimum. This reduces the voltage drop across the field winding.

iii) A high field flux is obtained by using low reluctance magnetic circuit.

iv) There is considerable sparking between the brushes and the commutator when the motor is used on A.C. supply. It is because the alternating flux establishes high currents in the coils short circuited by the brushes. When the short circuited coils break contact from the commutator, excessive sparking is produced. This can be eliminated by using high resistance leads to connect the coils to the commutator segments.

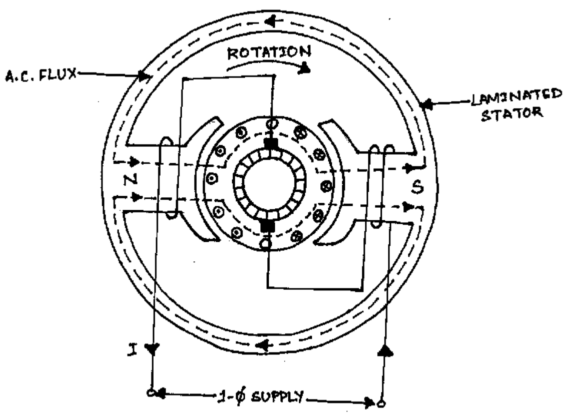
v) In order to reduce the effect of armature reaction thereby improving commutation and reducing armature reactance a compensating winding is used. This winding is put in the stator slot.

The drawback when A.C. supply is given to D.C. series motor (without modification) –

i) The efficiency is low due to hysteresis and eddy current loss.

ii) The power factor is low due to large reactance of the field and armature winding.

iii) The sparking at the brush is excessive.



COMMUTATOR MOTOR

Construction

The construction of an A.C. series motor is very similar to D.C. series motor except that above modification are incorporated.

This type of motor can be operated either on A.C. or D.C. supply and the resulting torque-speed curve is about the same in each case. For this reason it is sometime called universal motor.

Motors that can be used with a 1-phase A.C. source as well as a D.C. source of supply voltage are called universal motors.

Principle of Operation of A.C. series motor

When the motor is connected to an A.C. supply the same alternating current flows through the field and armature windings.

The field winding produces an alternating flux Φ that reacts with the current flowing in the armature to produce a torque.

Since both armature current and flux reverse simultaneously, the torque always acts in the same direction.

Characteristics of A.C. Series Motor

The operating characteristics are similar to those of D.C. series motor –

i) The speed increases to a high value with decrease in load.

ii) The motor torque is high for large armature current, thus giving high starting torque.

iii) At full load, the power factor is about 90 %, however at starting or when carrying overload power factor is low.

Application

The fractional horsepower A.C. series motor have high speed and large starting torque. Therefore be used to drive –

a) High speed vacuum cleaners.

b) Sewing Machine

c) Electric Shavers

d) Drills

e) Mechanical tools etc.

Repulsion Motor

A repulsion motor is similar to an A.C. series Motor except –

i) The brushes are not connected to supply but are short circuited. Hence current are induced in the armature conductor by transformer action.

ii) The field structure has non-silent pole construction

By adjusting the position of short circuited brushes on the commutator, the starting torque can be developed in the motor.

Construction

The field of the stator winding is connected to the 1 – Φ A.C. supply.

The armature or rotor with drum type winding like D.C. motor is connected to a commutator.

Here the brushes are not connected to the supply but are connected to each other or short circuited. Hence it is possible to vary the starting torque by changing the brush axis. So Commutator motor has better power factor than conventional 1-phase motor.

Principle of Operation

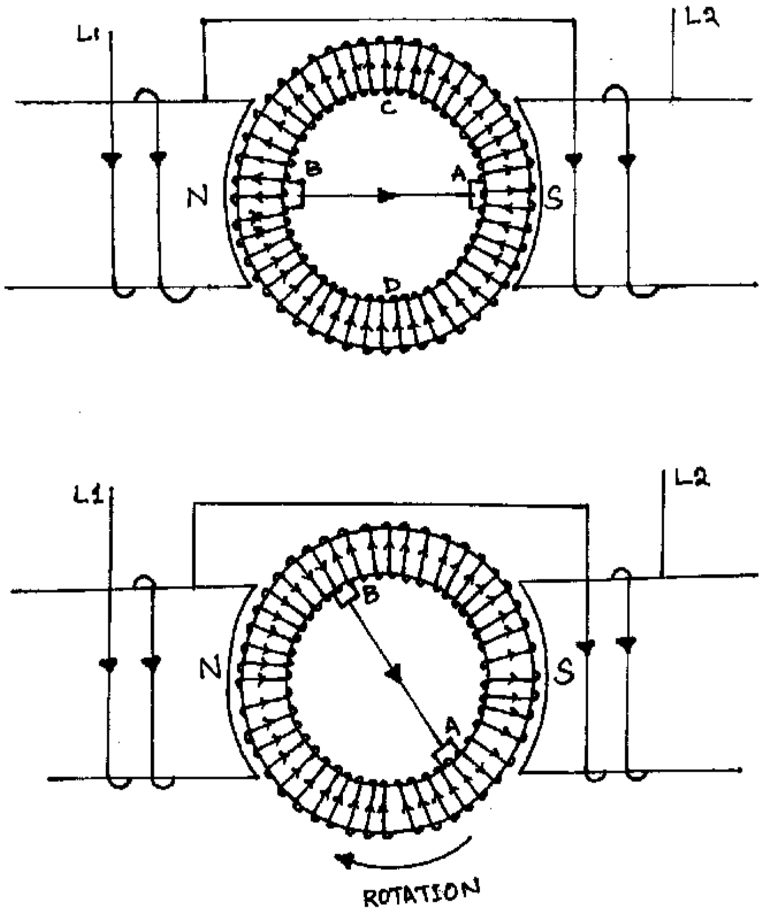
Fig. 1 shows two pole repulsion motor with its two short-circuited brushes

When field current is increasing in the direction shown the left hand pole is north pole and right hand pole is south pole.

i) Here the brush axis is parallel to the stator field.

When the stator winding is energized from 1 – Φ supply emf is induced in the armature conductor by induction. This emf will cause a current to flow in the armature conductor. By lens’s law the direction of the emf is such that magnetic field of the resulting armature current will oppose the increase in flux.

The current direction in armature conductor is shown in the Fig.



REPULSION MOTOR

With brushes set in this position, half of the armature conductors under the N-pole carry current inward and half carry current outward. The same is true under south pole.

So as much torque is developed in one direction as in the other and the armature remains stationary.

The armature will also remain stationary if the brush axis is perpendicular to the stator field axis as even then net torque is zero.

If the brush axis is at some angle other than 0^0 or 90^0 to the axis of stator field a net torque is developed on the rotor and rotor accelerate to it’s final speed.

Here in figure 2 because of the new brush position, the greater part of the conductor under the N-pole carry current in one direction. While the greater part of conductor under S-pole carry current in opposite direction.

With brushes in position 2 torque in developed in the clockwise direction and the rotor quickly attains the final speed.

The direction of rotation of the rotor depends upon the direction in which the brushes are shifted. If the brushes are shifted in clockwise direction from the stator field axis, the net torque acts in the clockwise direction and rotor accelerates in the clockwise direction and vice versa.

The total armature torque in a repulsion motor is

T_a = Sin 2α where α is the angle between brush axis and stator field axis.

For maximum torque, $2\alpha = 90^0$ or $\alpha = 45^0$. Thus adjusting α to 45^0 at starting, maximum torque can be obtained during starting period.

Characteristics

a) The repulsion motor has characteristics very similar to those of an A.C. series motor i.e. it has a high starting torque and a high speed at no load.

b) The speed which the repulsion motor develops for any given load will depend upon the position of the brushes.

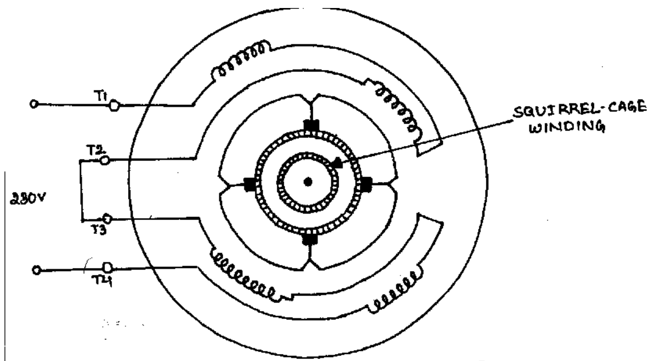
c) In comparison to other single phase motor, the repulsion motor has high starting torque and relatively low starting current.

Repulsion Induction Motor

The repulsion – Induction motor produces a high starting torque entirely due to repulsion motor action and when running, it function through a combination of Induction motor and repulsion motor action.

Construction

The Fig. shows the connection of a 4-pole repulsion Induction motor for 230 V operation. It consist of a stator and a Rotor.



REPULSION- INDUCTION MOTOR

i) The stator carries a single distributed winding fed from single-phase supply.

ii) The rotor is provided with two independent windings placed one side the other. The inner winding is a squirrel-cage winding with rotor bars permanently short circuited. The outer winding is a repulsion commutator armature winding placed over the squirrel cage winding.

The repulsion winding is connected to a commutator on which ride short circuited brushes.

Operation

When single phase stator winding is driven by an A.C. supply the repulsion winding is active. Consequently the motor starts as a repulsion motor with a corresponding high starting torque.

As the motor speed increases, the current shifts from the outer to inner winding due to the decreasing impedance of the inner winding with increasing speed. Consequently at running speed, the squirrel cage winding carries the greater part of rotor current. This shifting of repulsion motor action to induction motor action is thus achieved without any switching arrangement.

It may be seen that the motor starts as a repulsion motor. When running, it function through a combination of principle of induction and repulsion.

Characteristics

The no-load speed of a repulsion – Induction Motor is somewhat above the synchronous speed because of the effect of repulsion winding, however the speed at full load is slightly less than the synchronous speed in an induction motor.

The speed regulation of the motor is about 6 %.

The starting torque is 2.25 to 3 times the full load torque. The starting current is 3 to 4 times the full load current.

Application

This type of motor is used for applications requiring a high starting torque with essentially constant running speed.

Repulsion – Start Induction – Run motor

The action of repulsion motor is combined with that of a 1 – Φ induction motor to produce repulsion – start induction – run motor (also called Repulsion Start Motor)

This motor starts as an ordinary repulsion motor, but after it reaches about 75 % of its full speed, Centrifugal short – circuiting device / switch short circuits its commutator.

From then on it runs as an Induction Motor with a short – circuited squirrel – Cage Rotor.

After the commutator is short circuited, brushes do not carry any current, hence they may also be lifted from the commutator in order to avoid unnecessary wear and tear and friction losses.

Characteristics

The starting torque is 2.5 to 4.5 times the full load torque and the starting current is 3.75 times the full load value.

Due to their high starting torque, repulsion motors were used to operate devices such as refrigerators, pumps, compressor etc.

CHAPTER- VI

SPECIAL PURPOSE ELECTRIC MACHINES

INTRODUCTION

Special purpose electric machines have some features that distinguishes them from conventional machines.Stepper motor belongs to that type machine which rotates by a specific number of degrees in response to an input electrical signal and is widely used in digital control systems.

STEPPER MOTOR

Stepper motors are also known as stepping motors or step motors. A stepper motor is an electro-magnetic motor that rotates by a specific number of degrees in response to an input electrical signal. Typical step sizes are 2^0 ,2.5^0, 7.5^0, 15^0 for each electrical pulse. Note that there is no continuous energy conversion so that the rotor does not rotate continuously as in a conventional electric motor.The stepper motor converts electrical pulses into proportionate mechanical movement. Each revolution of stepper motor is made up of a series of definite individual steps. a step is defined as the angular rotation in degrees of the motor each time it receives the electrical pulse. such a step control is required in many applications. Figure 1.1 illustrates a simple application for a stepper motor. Each time the controller receives an input electrical signal, the paper is driven to a certain incremental distance. Stepper motors are relatively cheap and simple in construction and can be made to rotate in steps in either direction. These motors are excellent candidates for such applications as type-writers, control of floppy disc drives, numerical control of machine tools etc. The two most popular types of stepper motors are :

(i) Permanent-magnet (PM) Stepper Motor

(ii) Variable –reluctance(VR) Stepper Motor

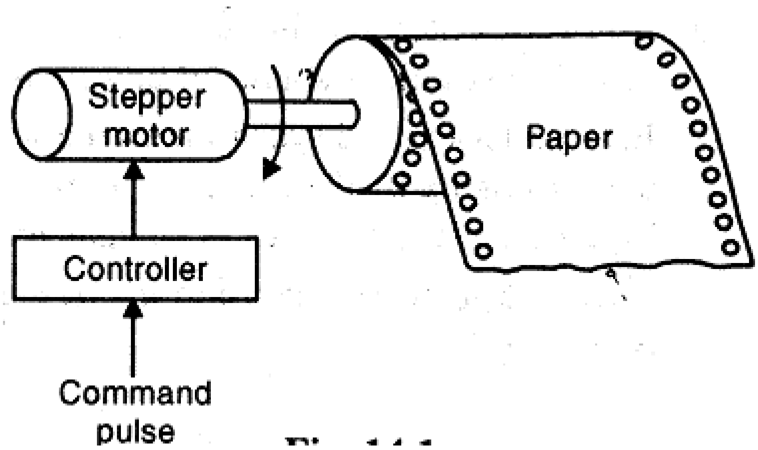


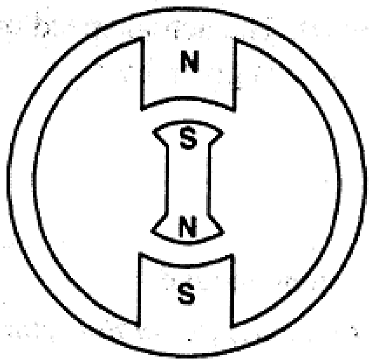
Fig. 1.1

The stator of a stepper motor of either type above carries stator windings which are energized from a dc source to create two or more stator poles. The stator poles are also called stator teeth. The rotor of a stepper motor may be a permanent magnet as in a Permanent Magnet stepper motor or a soft-iron material as in case of a variable reluctance motor. The rotor may also have two or more poles. The rotor poles are also called rotor teeth.

The stator coils are energized in groups referred to as phases. The stator windings may be 2-phase,3-phase or 4-phase windings. The phase windings are brought out to terminals for DC excitation .

PM Stepper Motor

The figure1.2 shows a two-pole 1-phase permanent magnet stepper motor. When the stator is energized, the excitation torque acts on the rotor. The rotor will move to a position where the excitation torque is zero i.e. the rotor will be aligned in parallel to the stator field.



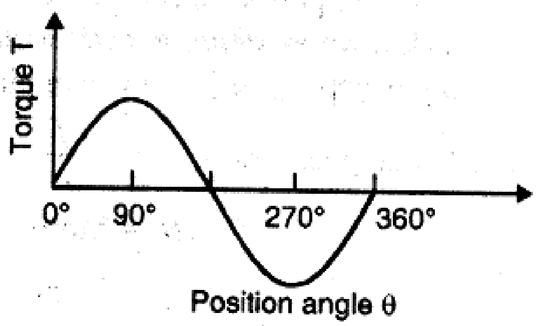


Fig.1.2

Fig.1.3

Fig 1.3 shows how excitation torque varies with thee rotor position for a PM rotor. Note that maximum torque is developed when the rotor is displaced from the stator field by either 90^0 or 270^0 . However , the torque is zero and the rotor is aligned (parallel) with the stator field.

(iii) VR Stepper Motor

Fig. 1.4 shows a 2-pole , single phase variable-reluctance(VR) stepper motor.When the stator is energized ,*reluctance torque* acts on the rotor (soft-iron material). The rotor will move a position where reluctance is minimum and air-gap flux is maximum. This means that rotor teeth will align with the energized stator poles.

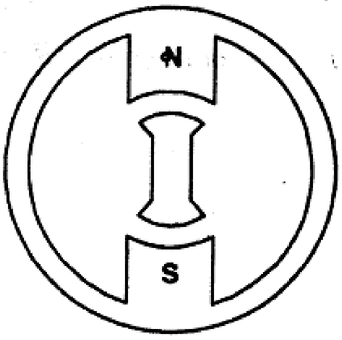


Fig.1.4

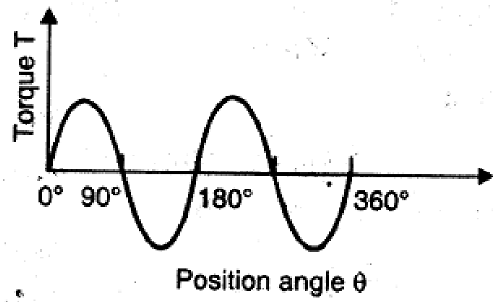


Fig.1.5

Fig.1.5 shows how reluctance torque varies with the rotor position for a VR soft-iron rotor.With the rotor at 0^0 or 90^0 , no torque is developed. Maximum torque is developed at 45^0 and 135^0 Which is the position where reluctance torque forces the rotor or move to position of minimum reluctance.

step angle :the angle through which the motor shaft rotates for each command pulse is called step angle. It can be shown that for any PM or VR stepper motor,the step angle can be found from the following two relation:

i) In terms of stator poles (N_s) and rotor poles (N_r), the step angle (α) is given by:

$$\text{Step angle, } \alpha = \frac{N_s - N_r}{N_s \times N_r} \times 360^0$$

where α= Step angle in degrees

(N_s)=Number of stator poles(or teeth)

(N_r)=Number of rotor poles (or teeth)

ii) In terms of stator phases (m) and rotor poles (N_r), the step angle is given by:

step angle,α= $\alpha=\frac{360^0}{m\,N_r}$

α= step angle in degrees

m=Number of stator phases

N_r=Number of rotor poles (or teeth)

stepping rate. An important specification of a stepper motor is the stepping rate. The number of steps per second is known as stepping frequency(f).The actual speed of a stepper motor depends on the step angle (α) and stepping frequency(f) and is given by :

Speed of stepper motor, N $=\frac{\alpha f}{6}$

N = motor speed in r.p.m.

f = stepping frequency i.e. steps/second

Example 1.1

Determine the step angle of a variable-reluctance stepper motor with 12 teeth in the stator and 8 rotor teeth.

Solution :

Number of stator teeth, N_s = 12

Number of rotor teeth, N_r = 8

Step angle, $\alpha=\frac{N_s-N_r}{N_s \times N_r} \times 360^0 = \frac{(12-8)}{(12\times 8)} \times 360^0 = 15^0$/step

Example 1.2

A stepper motor has a step angle of 10^0 and is required to rotate at 200 r.p.m. Determine the pulse rate(steps/second) for this motor.

Solution :

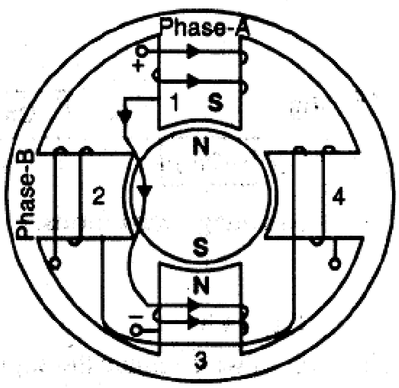
motor speed, N $=\frac{\alpha f}{6}$

Hence , Pulse rate(steps per second) for this motor = $\frac{6\times N}{\alpha} = \frac{6\times 200}{10}$ = 120 steps/second

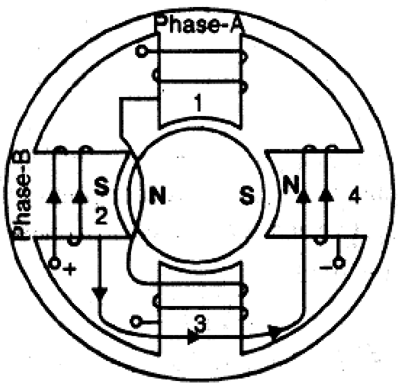
PERMANENT –MAGNET (PM) STEPPER MOTOR

A permanent-magnet(PM) stepper motor is a popular type of stepper motor.It operates on the principle of interaction between permanent-magnet and electromagnetic field.

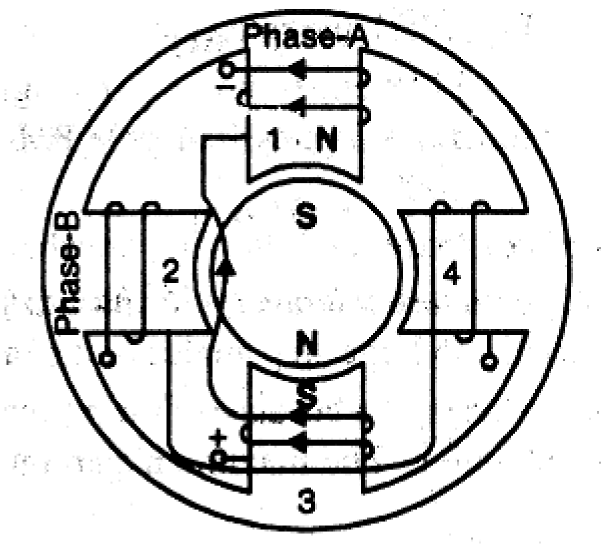
CONSTRUCTION : The stator construction of a PM stepper motor is composed of steel laminations and carries stator windings. The stator phase windings are energized from a d.c. source to create two or more stator poles. The rotor of the motor is a permanent-magnet made up of high retentivity steel alloy.The rotor has even number of poles. Fig.1.6 shows a two-phase,2-pole PM stepper motor. The motor has two rotor poles.The stator coils are grouped to form 2-phase winding i.e.phase-A winding and phase-B winding.The phase winding terminals are brought out for d.c. excitation.



(i)



(ii)



(iii)

Fig.1.6

OPERATION : for this PM stepper motor,the number of rotor poles,N_r = 2 and number of phases, m=2.

Step angle,α =mN_r=360^0/(2x2)= 90^0/step

(i) When only phase-A winding is energized by a constant current as shown in Fig.1.6(i) stator tooth 1 becomes the south pole. This makes the north pole of the PM rotor to align parallel with the south pole(stator tooth 1) of the stator. The rotor will remain locked in this position as long as phase-A winding remains energized. The first row of truth table in Fig. shows that only phase-A winding is excited while phase-B winding is unexcited. Under this condition, step angle α =0^0.The applied voltage waveforms in Fig also tally with the facts shown in the truth table.

(ii) If phase A winding is de-energized and phase-B winding is energised as shown in Fig.1.6(ii), stator tooth 2 becomes south pole. As a result, the north pole of the PM rotor aligns parallel with the south pole(stator tooth2) of the stator. Thus the rotor has displaced 90^0 in the anticlockwise direction.

(iii) If phase B winding is de-energized and phase-A winding is excited by a reverse current the rotor will further rotate 90^0 in anticlockwise direction as shown in Fig1.6(iii). Now the north pole of PM motor aligns with the stator tooth 3.

Truth Table

Cycle	Phase A	Phase B	Position $\delta°$
+	1	0	0
	0	1	90
−	−1	0	180
	0	−1	270
+	1	0	360

Fig.1.7

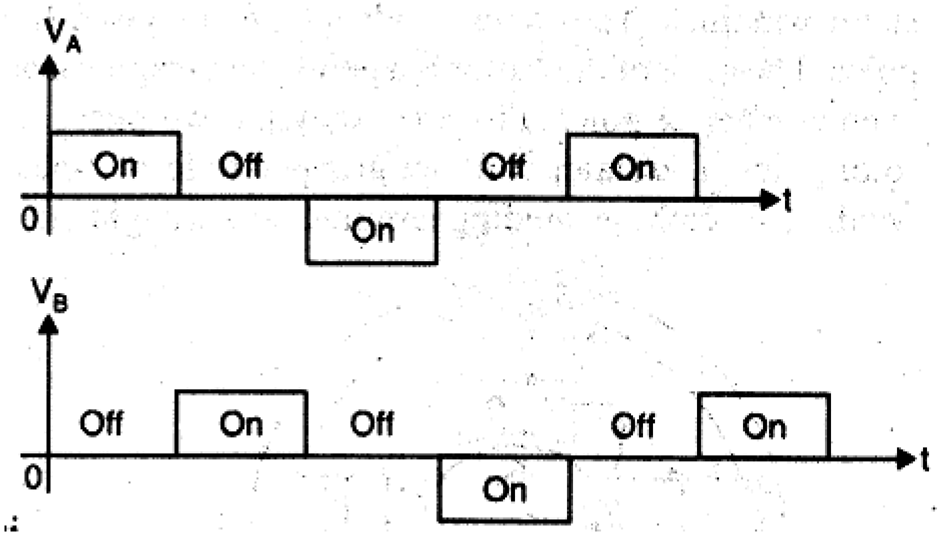


Fig.1.8

(iv) So far the rotor has completed one-half revolution. However, if we continue the appropriate switching the rotor will complete one revolution in 90^0 steps.

We can change the step angle α of a PM stepper motor by changing the number of rotor poles N_r and the number of phases(m).Thus for a 3-phase,24-pole PM stepper motor, the step angle $\alpha = 360^0/mNr = 360^0/3x24 = 5^0$/step.

Limitations : The PM stepper motor has the following draw\backs :

i) It is difficult to make a small permanent magnet rotor with a large number of poles. Therefore ,PM stepper motors are restricted to large step angles in the range of 30^0 to 90^0.

ii) The PM stepper motors have high inertia because of the permanent-magnet rotor. Therefore, these motors have slow acceleration. the maximum step rate (Stepping frequency) is 300 steps/second.

iii) The PM stepper motors have high rotational speed because of large stepping angle. Therefore, motor torque for a given output power is low.

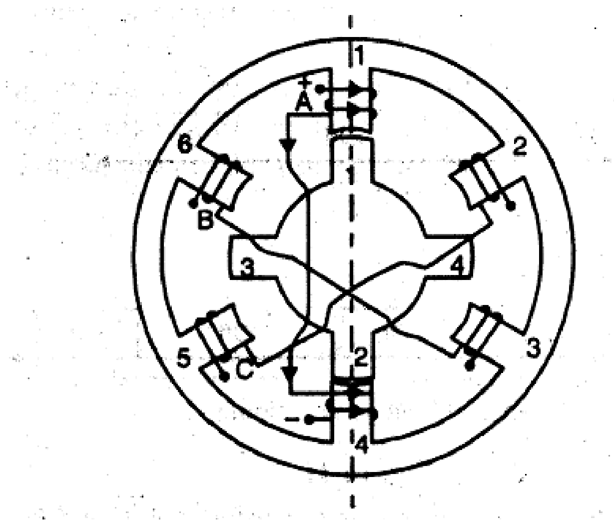
VARIABLE RELUCTANCE(VR) STEPPER MOTOR

The variable Reluctance stepper motor(VR) stepper motor operates on the same principle as the reluctance motor. that is, when a piece of ferro-magnetic material is free to rotate and is placed in a magnetic field the torque acts on the material to bring it to the position of minimum reluctance to the path of magnetic flux.

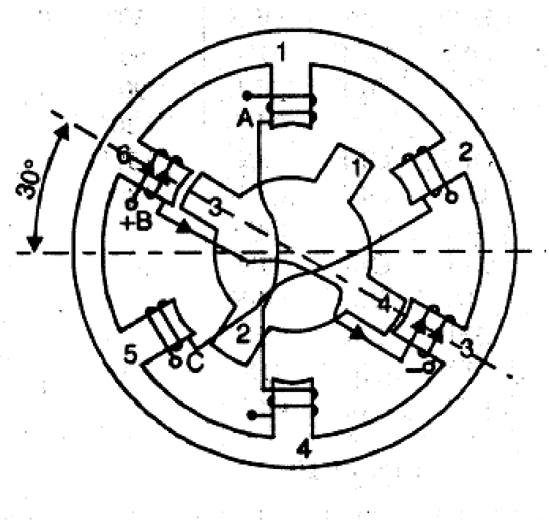
CONSTRUCTION : The stator construction of a VR stepper motor is the same as that of a PM stepper motor. The stator phase windings are wound on each stator tooth. The rotor is made of soft steel with teeth and slots .Figure shows the basic Variable-Reluctance stepper motor. In this circuit, the rotor is shown with fewer teeth than stator. This ensures that only one set of stator and rotor teeth will align at any given instant. In Fig. the stator has six teeth and the rotor has four teeth. The stator has three phases – A,B and C with teeth 1 and 4, 3 and 6 and 2 and 5 respectfully .For this VR stepper motor,

step angle,α $= \frac{N_s - N_r}{N_s \times N_r} \times 360^0 = \frac{6-4}{6\times4} \times 360^0 = 30^0$/step

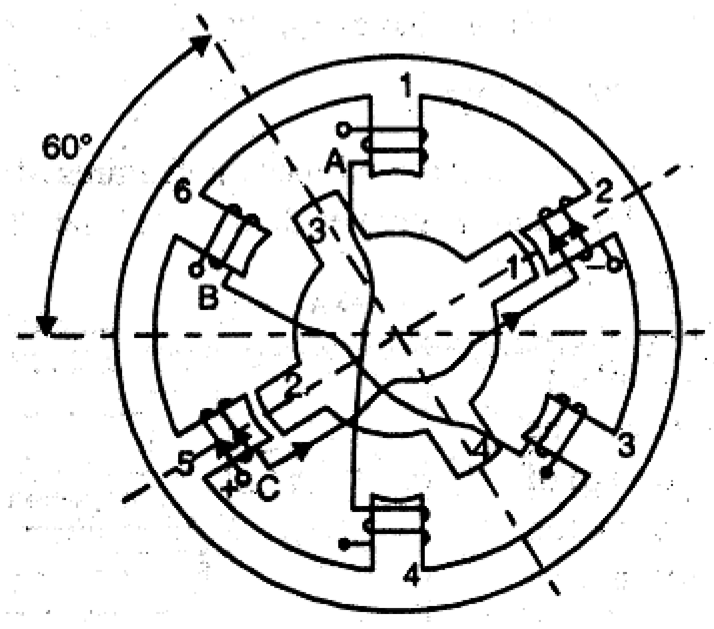
Therefore, the rotor will turn 30^0 each time a pulse is applied.



(i)



(ii)



(iii)

Fig.1.9

OPERATION : When the phase winding is energized,the rotor teeth will align with the energized stator poles.

i) Fig.1.9(i) shows the position of the rotor when phase A is energized with a constant current.As long as phase A is energized,the rotor will be held stationary.Note that in this condition,the rotor teeth 1 and 2 are aligned with the energized stator teeth 1 and 4.the step angle $\alpha=0^0$.Also refer to truth table and applied voltage waveform.

ii) when phase A is switched off and phase B is energized,the rotor will turn 30^0 clockwise so that the rotor teeth 3 and 4 align with the energized stator teeth 6 and 3.

iii) The effect of de-energising phase B and energizing phase C is shown in Fig.1.9(iii).In this circuit,the rotor has further moved 30^0 clockwise so that rotor teeth 1 and 2 align with energized stator teeth 2 and 5.

iv) after the rotor has displaced 60^0 clockwise from iots starting point,the step sequence has completed one cycle. The truth table in fig. shows the switching sequence to complete a full 360^0 rotation for the motor with six stator poles and four rotor poles.

Truth Table

Cycle	Phase A	Phase B	Phase C	Position
1	ON	OFF	OFF	0°
	OFF	ON	OFF	30°
	OFF	OFF	ON	60°
2	ON	OFF	OFF	90°
	OFF	ON	OFF	120°
	OFF	OFF	ON	150°
3	ON	OFF	OFF	180°
	OFF	ON	OFF	210°
	OFF	OFF	ON	240°
4	ON	OFF	OFF	270°
	OFF	ON	OFF	300°
	OFF	OFF	ON	330°
5	ON	OFF	OFF	360°

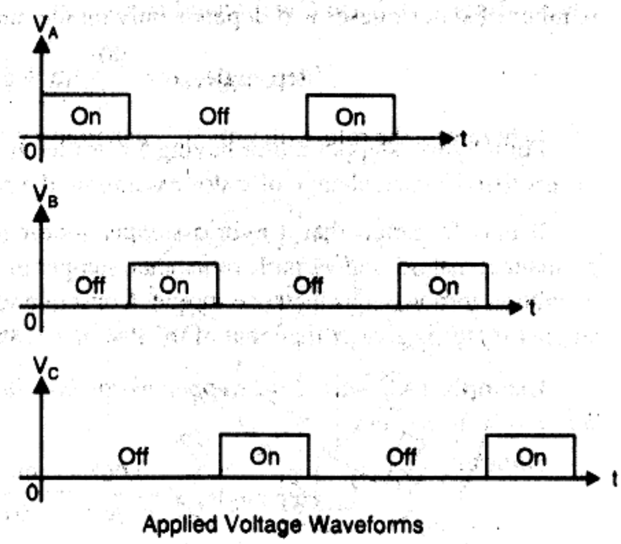


Fig.1.10 Fig.1.11

The direction of rotation will be reversed if the switching sequence is in the order of A,C and B.For this particular motor,applied voltage must have at least five cycle for one revolution.

HYBRID STEPPER MOTOR

The hybrid stepper motor combines the features of the PM and the VR stepper motors.The torque developed by this motor is greater than that of the PM or VR stepper motor.

Construction : Fig1.12 shows the basic construction of a hybrid stepper motor.The stator construction is similar to that of a VR or PM stepper motor. However, the rotor construction combines the design of the rotors of a VR and a PM stepper motor. The rotor of a hybrid stepper motor consists of two identical stacks of soft iron as well as an axially magnetized round permanent magnet. Soft iron stacks are attached to the north and south poles of the permanent magnet as shown in Fig.1.12

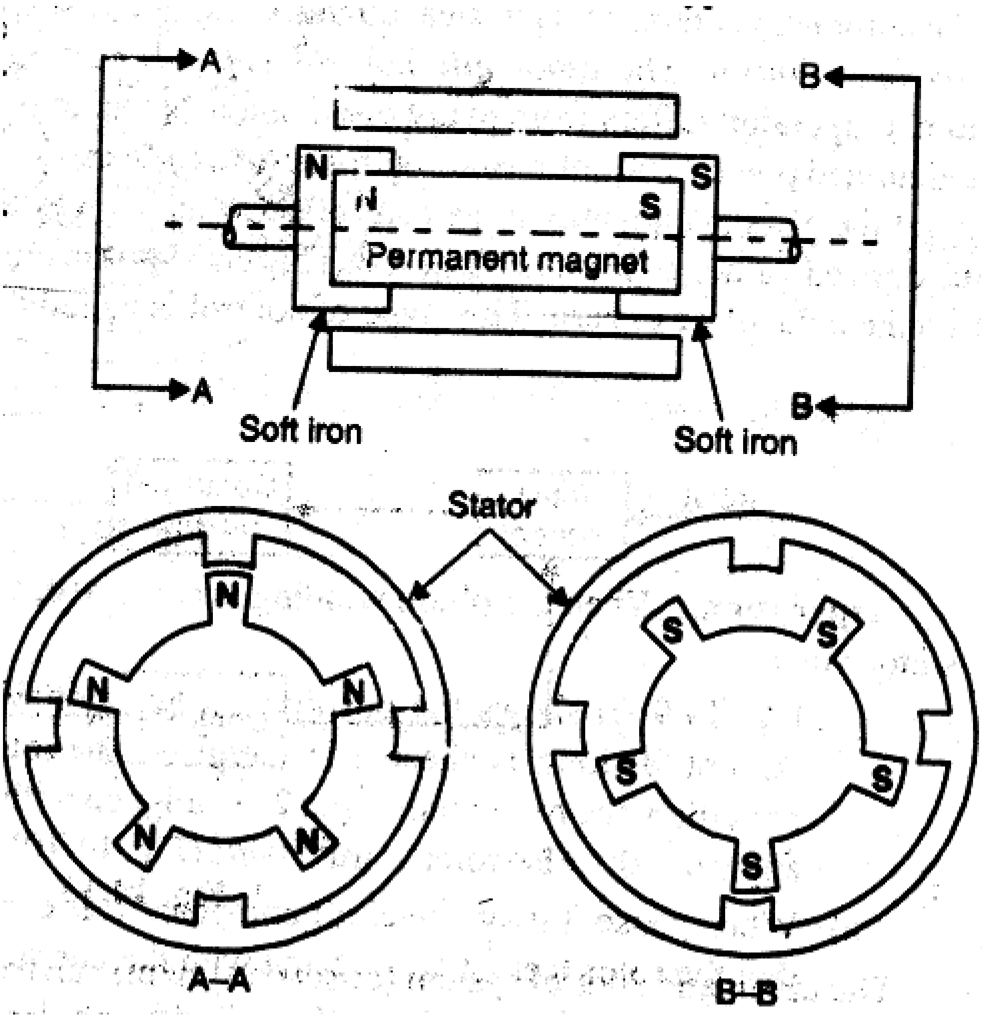


Fig 1.12

The rotor teeth are machined on the sot iron stacks. Thus the rotor teeth on one end become the north pole and those at the other end become the south pole.

This rotor teeth of both north and south poles are displaced in angle for the proper alignment of the rotor pole with that of the stator as shown in Fig.1.12

OPERATION : The operating mode of the hybrid stepper motor is very similar to that of a PM or VR stepper motor.The phase windings are energized in proper sequence and the rotor rotaes in steps.Unlike the VR or PM stepper motors,the step angle of a hybrid stepper motor is independent of the number of stator phases and depends only on the number of rotor teeth(N_r).It is given by :

Step angle, α = $90^0/N_r$ in deg

For a hybrid stepper motor having 5 rotor teeth,the step angle alpha =$90^0/N_r$ =90^0/5 =18^0/step. It means that for each change of stator excitation,the rotor will turn by a step of 18^0.

It may be noted that a hybrid stepper motor operates under the combined principles of the PM and VR stepper motors.Therefore,the hybrid motor develops both excitation torqe and relucatance torque.Consequently the resultant torque developed by the hybrid stepper motor is greater than that of the PM or VR stepper motor.

…………………………………………..xxxxxxxxxxxxxxx………………………………………

CHAPTER VII

THREE PHASE TRANSFORMER

All alternating current electrical energy is nearly generated by three phase alternating current generators. Similarly three phase systems are used for transmission and distribution of electrical energy. There are several reasons why a three phase system is preferred over a single phase system. Some of the important reasons are

- Smaller size - KVA ratings of three phase generators and horse power ratings of three phase motors for a given physical size are higher than those of similar single phase units.
- Superior operating characteristics - operating characteristics of three phase motors and other appliances are superior to those of similar single phase units.
- Better efficiency - the efficiency of transmission and distribution of power in three phase system are better than in a single phase system.

Alternating current generated through a three phase generator has to be transmitted at higher voltage level for economic reason. Again at the receiving end of transmission line it is necessary to transform the energy through a suitable lower voltage level for distribution. It is therefore often necessary to transform the three phase voltage system to a higher or lower value.

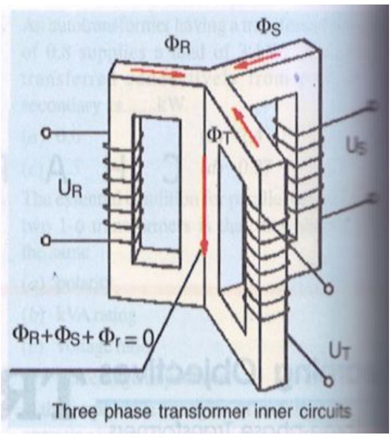
Electric energy may be transferred from one three phase current to another three phase current with a change in voltage by means of a three phase transformer. Voltage transmission on a three phase system may also be performed by using three separate single phase transformer with the winding of the transformer connected in star or delta.

Advantages of single three phase transformer over a bank of three single phase transformers

Recently, three phase transformer are increasingly being used for both step up and step down applications for the following reasons-

- The cost of one three phase transformer is less than the cost of three single phase transformer required to supply the same KVA output.
- The3 phase transformer weights less and occupies less space than 3 single phase transformer.
- The bus bar structure, switchgear and other wiring for a three phase transformer installation
  are simpler than those for three single phase transformer.

  But there is one major advantage in using a bank of three single phase transformers than a
  Single three phase transformer. If one single phase transformer among the bank becomes defective, it can be disconnected and power can be supplied by the other two single phase transformers unless replacement/repair is possible. However in a three phase transformer, If one of the phase winding becomes defective, the entire transformer must be taken out of a Service for repair work, thereby completely disturbing the power supply.



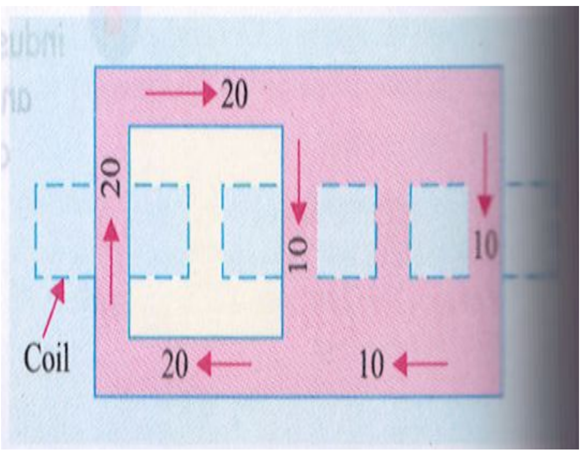


Fig.1.1 Fig.1.2

1.2. Construction

The three phase transformers are also core type and shell type. The basic principle of a three phase transformer is shown in figure.1.1, in which only primary windings have been shown interconnected in star and put across three phase supply. Three cores are 120° apart and their empty legs are shown contact with each other. The centre leg formed by these three carriers the flux produced by the three phase currents I_R, I_Y and I_B. As at any instant **$I_R + I_Y + I_B = 0$**, hence the sum of three fluxes is also zero. Therefore it will make no difference if the common leg is removed. In that case any two legs will act as their return path for the third Just as in a three phase system any two conductors act as the return for the current in the third conductor.

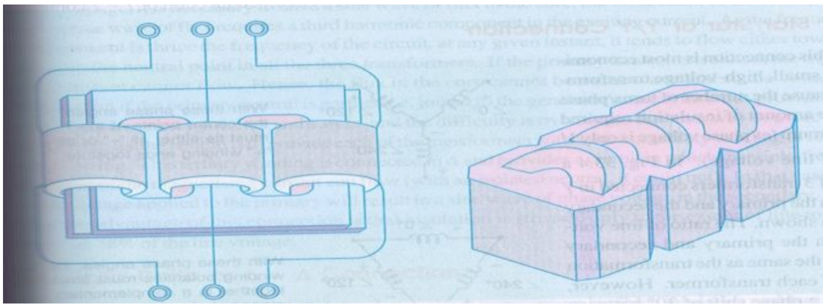


Fig.1.3

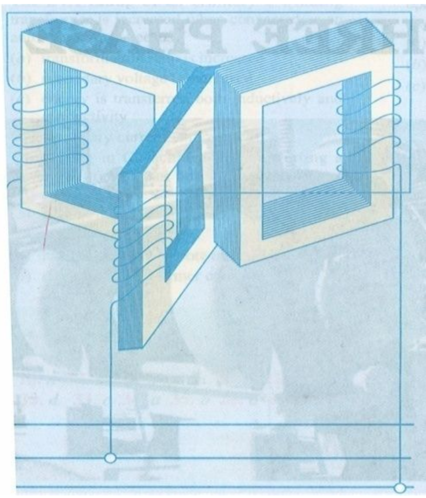


Fig.1.4

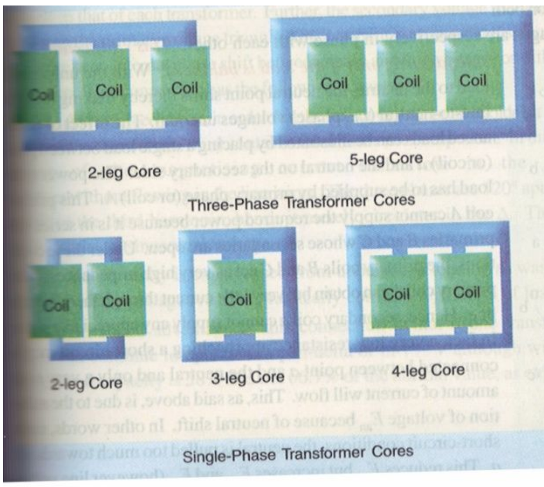


Fig.1.5

1.3. Grouping of the three phase transformer

Three phase transformers are divided into four groups according to their phase displacement between the line voltage on the hv and lv side.

Group1- 0 degree displacement (star-star or delta-delta)

Group2 - 180 degree displacement (star-star or delta-delta but the secondary is reversed)

Group3- +30 degree displacement (sta-delta)

Group4- -30 degree displacement (delta-star)

Thus a connection Yd11 gives the following information

Y indicates that hv is connected in star

d indicates that lv is connected in delta

11 indicate that lv line voltage lags hv line voltage by +30 degree. (Measured from hv phasor in anticlockwise direction).

The phase difference between the hv & lv windings for different types of connection can be represented by comparing it with the hour hand of the clock. When the hour hand of the clock is at 12 O'clock position, the phase displacement is zero. Similarly

Position of hour hand of clock	Phase displacement
0	0°
11	+30°
1	-30°
6	180°

Depending on the phase displacement of the voltages of hv (high voltage) & lv (low voltage) sides, transformers are classified into groups called "Vector group". Transformer having the same phase displacement between the hv & lv sides are classified into one same group. For successful parallel operation of transformers, they should belong to the same vector group. For example, a star-star connected three phase transformer can be paralleled with another three phase transformer whose windings are either star-star connected or delta-delta connected. A star-star connected transformer cannot be paralleled with another star-delta connected transformer as this may result in short-circuiting of the secondary side.

1.4. Three phase transformer connection

There are various methods available for transforming three phase voltages to higher or lower 3 phase voltages i.e. For handling a considerable amount of power. Usually star connection is used for high voltage transformation and delta connection is used for high current transformation. The most common connection are

1. Y-Y
2. Δ-Δ
3. Y-Δ
4. Δ-Y
5. Open Δ or V-V
6. Scott connection or T-T connection

1.5. Star/Star or Y-Y connection:-

This connection is most economical for small, high voltage transformer because the no of turns per phase and the amount of insulation required is minimum (as phase voltage is only 1/√3 of line

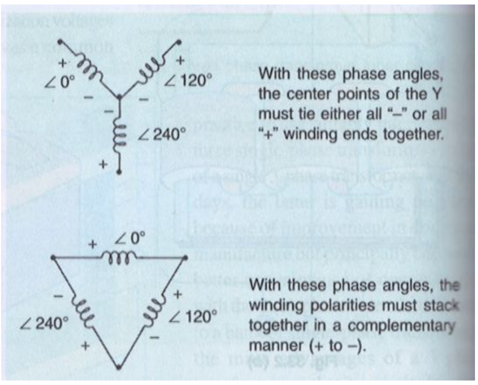


Fig.1.6

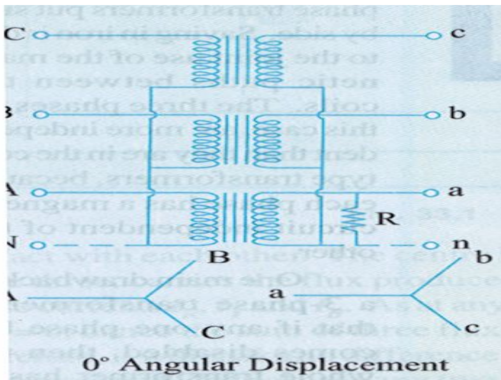


Fig.1.7

Voltage). In figure1.7 a bank of three transformers connected in star on both the primary and secondary sides are shown. The ratio of line voltage on the primary and secondary sides is the same as the transformation ratio of each transformer. However there is a phase sift of 30° between the phase voltages and line voltages both on the primary and secondary sides. Of course line voltages on both sides as well as primary voltages are respectively in phase with each other. This connections works satisfactorily only if the load is balanced. With the unbalanced load to the neutral, the neutral point shifts there by making the 3 line- to-neutral

(I.e. phase) voltages unequal. The effect of unbalanced loads can be illustrated by placing a single load between phase (or coil) a and the neutral on secondary side. The power in the load has to be supplied by primary phase (or coil) A. This primary coil A cannot supply the required power because it is in series with primaries B and C whose secondaries are opened. Under these condition the primary coils B and C act as very high impedances so that primary coil A can obtain but very little current through them from the line. Hence secondary coil a cannot supply appreciable power. In fact, a very low resistance approaching a short circuit may be connected between point A and the neutral and only a very small amount of current will flow. This, as said above, is due to the reduction of voltage E_{an} because of neutral shift. In other words, under short-circuit condition, the neutral is pulled too much towards coil a. This reduces Ean but increases Ebn & Ecn (however line voltage E_{AB}, E_{BC},E_{CA} are unaffected). On the primary side, Ean will be practically reduced to zero whereas E_{BN} & E_{CN} will rise to nearly full primary line voltage. This difficulty of shifting (or floating) neutral can be obviated by connecting the primary neutral (shown dotted in the figure) back to the generator so that primary coil A can take its required power from between its line and the neutral. It should be noted that if a single phase load is connected between the lines a and b, there will be a similar but less pronounced neutral sheet which results in an over voltage on one or more transformers.

Another advantage of stabilizing the primary neutral by connecting it to neutral of the generator is that it eliminates distortion in the secondary phase voltages. This is explained as follows. For delivering a sine wave of voltage, it is necessary to have a sine wave of flux in the core, but on account of the characteristics of iron, a sine wave flux requires a third harmonic component in the exciting current. As the frequency of this component is thrice the frequency of three circuit, at any given instant of time, it needs it tends to flow either towards or away from the neutral point in all the three transformers. If the primary neutral is isolated the triple frequency current cannot flow. Hence, the flux in the core cannot be a sine wave and so the voltages are distorted. But if the primary neutral is earthed i.e. joined to the generator neutral, then this provides a path for the triple frequency currents and e.m.fs and the difficulty is overcome. Another way of avoiding this trouble of oscillating neutral is to provide each of the transformers with a third or tertiary winding of relatively low KVA rating. This tertiary winding connected in delta and provides a circuit in which the triple frequency component of the magnetising current can flow (with an isolated neutral, it could not). In this case a sine wave of voltage applied to the primary will result in a sine wave of phase voltage in the secondary. As said above, the advantage of this connection is that insulation is stressed only to the extent of line to neutral voltage i.e. 58% of the line voltage.

1.6. Delta-Delta or Δ-Δ connection:-

This connection is economical for large, low voltage transformers in which insulation problem is not so urgent, because it increases the number of turns/phase. The transformers connection and voltage triangles are shown in fig 1.8 The ratio of transformation between primary and secondary line voltage is exactly the same as that of each transformers. Further, the secondary voltage triangle abc occupy the same relative position as the primary voltage triangle ABC i.e. there is no angular displacement between the two. Moreover, there is no internal phase shift between phase and line voltages on either side as was the case in Y-Y connection. This connection has the following advantages:

-4-

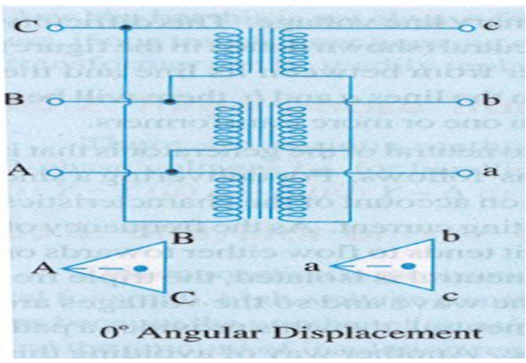


Fig.1.8

1. As explained above, in order that the output voltage be sinusoidal, it is necessary that the magnetising current of the transformer must contained a third harmonic component. In this case third harmonic component of the magnetising current can flow in the Δ connected transformer primaries without flowing in the line wires. The three phases are 120° apart which is 3X120°=360° with respect to the third harmonic, hence it merely circulates in the Δ. Therefore the flux is sinusoidal which results in sinusoidal voltages.

2. No difficulty is experienced from unbalanced loading as was the case in Y-Y connection. The three phase voltages remain practically constant regardless of load imbalance.

3. An added advantage of this connection is that if one transformer becomes disable, the system can continue to operate in open delta or in V-V although with reduced available capacity. The reduced capacity is 58% and not 66.7% of the normal value as explained in Art.1.9.

1.7. Wye/Delta or Y-Δ connection:-

The main use of this connection is at the substation end of the transmission line where the voltage is to be stepped down. The primary winding is Y connected with grounded neutral as shown in fig1.9 the ratio between the secondary and primary line voltage is 1/√3 times the transformation ratio of each transformer. There is a 30° shift between the primary and secondary line voltages which means that a Y-Δ transformer bank cannot be paralleled with either a Y-Y and Δ-Δ bank. Also, a third harmonic current flows in the Δ to provide a sinusoidal flux.

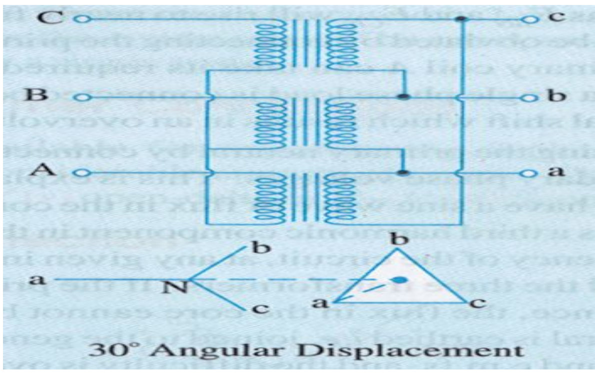


Fig.1.9

1.8. Delta/Wye or Δ-Y connection:-

This connection is generally employed where it is necessary to step up the voltage as for example at the beginning of high tension transmission system. The connection is show in fig1.10 the neutral of the secondary is grounded for providing three phase four wire service. In recent years, these connections has gained considerable popularity because it can be used to serve both the three phase power equipment and single phase lighting circuit.

This connection is not open to the objection of a floating neutral and voltage distortion because the existence of a Δ connection allows a path for the third harmonic currents. It would be observed that the primary and secondary line voltages and line currents are out of phase with each other by 30°. Because of this 30° shift it is impossible to parallel such a bank with a Δ-Δ and **Y-Y** bank of transformers even though the voltage ratios are correctly adjusted. The ratio of secondary to primary voltage is √3 times the transformation ratio of each transformer.

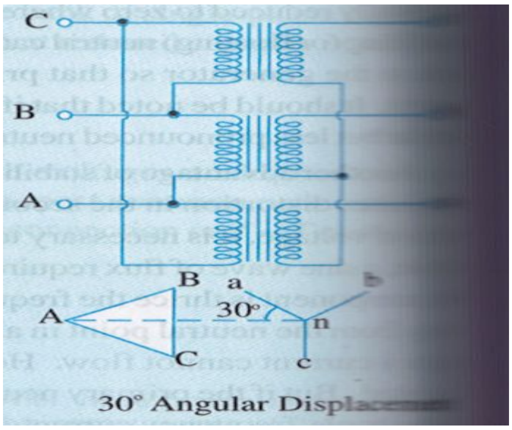


Fig.1.10

Example 1.1. A 3 phase, 50 Hz transformer has a delta-connected primary and star connected secondary, the line voltage being 22000 V and 400V respectively. The secondary has a star connected balanced load at 0.8 power factor lagging. The line current on the primary side is 5A.Determine the current in each coil of the primary and in each secondary line. What is the output of the transformer in KW ?

Solution : It sould be noted that in 3 phase tranformer, the phase tranformation ratio is equal to the turn ratio but the terminal or line voltages depend upon the method of connection employed. The delta/star connection is shown in figure 1.11 .

Phase voltage on primary side= 22000V

Phase voltage on secondary side= 400/√3

K=400/22000x√3 = 1/55√3

Primary phase current = 5/√3 A

Secondary phase current= (5/√3)/K =(5/√3)/(1/55√3) =**275 A**

Output = $\sqrt{3}V_LI_L\cos\Phi = \sqrt{3}x400x275x0.8 =$ **15.24 KW**

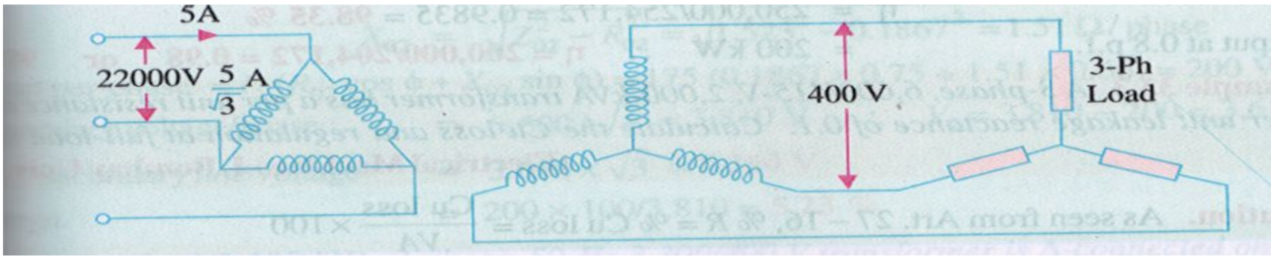


Fig.1.11

Example 1.2. A 500KVA, 3 phase, 50 Hz transformer has a voltage ratio (line voltage) of 33/11 KV and is delta/star connected. The resistances per phase are: high voltage 35 Ω, low voltage 0.876 Ω and the iron loss is 3050 W. calculate the value of efficiency at full load and ½ of full laod respectively A) at unity P.F. and B) 0.8 P.F.

Solution: Transformation ratio (K)= 11000/√3x33000 = 1/3√3

Per phase R_{02}= 0.876+(1/3√3)^{2}x35=2.172 Ω

Secondary phase current= 500000/(√3x11000) = 500/11√3 A

Full load condition :

Full load total Cu loss= 3x(500/11√3)^{2}x2.172 =4490W

Iron loss=3050 W

Total full load losses=4490+3050 =7540 W:

Outout at unity P.F= 500KW

Full load efficiency=500000/507540=0.9854 or **98.54 %**

Output at 0.8 P.F=0.8x500=400KW

Efficincy=400000/407540= 0.982 or **98.2%**

Half load condition :

Output at unity P.F= 250 KW

Cu losses = (1/2)^{2}x4490 =**1,222 W**

Total losses= 3050+ 1222= 4172 W

Efficiency=250000/254172=0.9835 or **98.35%**

Output at 0.8 P.F=200 KW

Efficiency= 200000/204172= 0.98 or 98%

1.9. Open- Delta or V-V Connection.

If one of the transformers of a Δ-Δ is removed and 3phase supply is connected to the primaries as shown in Fig. 1.12, then three equal 3 phase valtages will be at the secondary terminals on no load. This method of transforming 3-pahse power by means of only two transformers is called the open -Δ or V-V connection.

It is employed:

1. When the three-phase load is too small to warrant the installation of full three phase transformer bank.
2. When one of the transformers in a Δ-Δ bank is disabled, so that service is continued although at reduced capacity, till the faulty transformers is repaired or a new one is substituted.
3. When it is anticipated that in future the load will increase necessitating the closing of open delta.

One important point to note is that the total load that can be carried by a V-V bank is not two-third of the capacity of a Δ-Δ bank but it is only 57.7% of it. That is a reduction of 15% (STRICTLY, 15.5%) from its normal rating. Suppose there is Δ-Δ bank of three 10-kVA transformers. When one transformer removed, then it runs in V-V. The total rating of the transformer kVA rating but only 0.866 of it i.e. 20x0.866=17.32 (or 30x0.57=17.3kVA). The fact that the ratio of V- capacity to Δ-capacity is 1/ $\sqrt{3} = 57.7\%$(or nearly 58%) instead of 66.67 percent can be proved as follows:

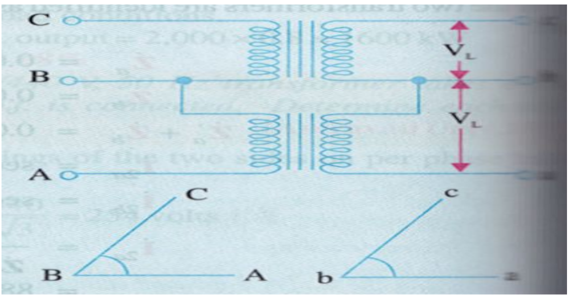


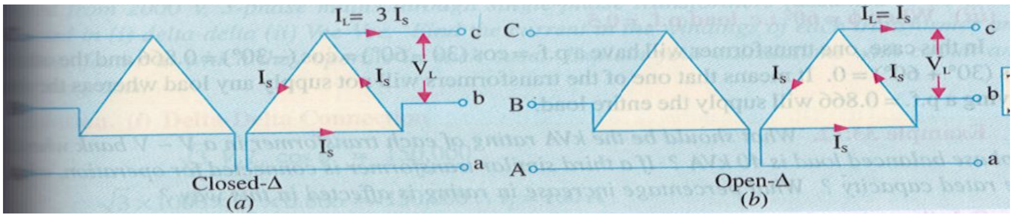
Fig1.12

As seen from fig 1.13(a)

Δ-Δ capacity = $\sqrt{3}.V_L I_L = \sqrt{3}.V_L(\sqrt{3}I_S) = 3\,V_L.I_S$

In Fig1.13 (b) it is obvious that when Δ-Δ bank becomes V-V bank, the secondary line current I_L becomes equal to the secondary phase current I_S.

(V-V- capacity/Δ-Δ capacity) = $\sqrt{3}.V_L I_S / 3\,V_L.I_S = 1/\sqrt{3} = 0.577$ or 58%



(a) Fig1.13 (b)

It means that the 3-phase load which can be carried without exceeding the rating of the transformers is 57.7 per cent of the original load rather than the expected 66.7%.It is obvious from above that when one transformer is removed from a Δ-Δ bank.

1. The bank capacity is reduced from 30 kVA to 30X0.577=17.3KVA and not to 20 kVA as might be tought off hand.
2. Only 86.6% of the rated capacity of the two remaining transformers is avalible (i.e. 20X0.866=17.3kVA). In other words, ratio of operating capacity to avalible capacity on an open -Δ is 0.866. This factor of 0.866 is sometimes called the utility factor.
3. Each transformer will supply 57.7% of load and not 50% when operating in V-V.

However, it is worth noting that if three transformers in a Δ-Δ bank are delivering their rated load and one transformer is removed, the overload on each of the remaining transformers is 73.2% because

(Total load in V-V)/ (VA/transformer) = $\sqrt{3}.V_L I_S / V_L I_S = \sqrt{3} = 1.732$

This over-load may be carried temporarily but some provision must be made to reduce the load if overheating and consequent breakdown of the remaining two transformers is to be avoided.

The disadvantages of this connection are:

1. The average power factor at which the V-bank opeates is less than that of the load, this power factor is actually 86.6% of the balanced load power factor. Another significant point to note is that, except for a balanced unity power factor load, the two transformers in the V-V bank operate at different power factors.
2. Secondary terminal voltages tend to become unbalanced to a great extent when the load is increased, this happens even when the load is perfectly balanced.

   It may, however be noted that if two transformers are operating in V-V and loaded to rated capacity in the above example, to 17.3kVA, the addition of a third transformer increases the total capacity by √3 or 173.2% (i.e to 30kVA). it means that for an increase in cost of 50% for the third transformer. The increase in capacity is 73.2%when converting from a V-V system to a Δ-Δ system.

1.10. Power supplied by V-V connection:

When a V-V bank of two transformer supplies a balanced 3-phae load of power factor cos ϕ, then one transformer operates at a p.f. of $\cos(30^0 - \phi)$ and the other at $\cos(30^0 + \phi)$. Consequently, the two transformers will not have the same voltage regulation.

$P_1 = KVA \times \cos(30^0 - \phi)$ And $P_2 = KVA \times \cos(30^0 + \phi)$

i) When $\phi=0$ i.e. load p.f. =1

Each transformer will have a p.f. = $\cos 30^0 = 0.866$

ii) When $\phi=30^0$ i.e. load p.f. =0.866,

In this case, one transformer has a p.f. of $\cos(30^0 - 30^0) = 1$ and the other of $\cos(30^0 + 30^0) = 0.866$

iii)) when $\phi=60^0$ i.e. load p.f.=0.5, In this case, one transformer has a p.f. of cos $(30^0 - 60^0)$ = cos (-30^0)=0.866 and the other of cos $(30^0 + 60^0)$ = cos (90^0)=0. It means that one of the transformers will not supply any load whereas the other having a power factor of 0.866 will supply the entire load.

Example 1.3. What should be the kVA rating of each transformer in a V – V bank when the 3 – phase balanced load is 40 kVA? If a third similar transformer is connected for operation ,what is the rated capacity? What percentage increase in rating is affected in this way?

Solution. As pointed out earlier, the kVA rating of each transformer has to be 15% greater.

kVA / transformer = (40 / 2) x 1.15 = **23**

Δ – Δ bank rating = 23 x 3 = 69; Increase = [(69 – 40) / 40] x 100 = **72.5 %**.

1.11. Scott Connection or T-T connection:

This is a connection by which 3-phase to 3-phase transformation is accomplished with the help of two transformers as shown in Fig. 1.14. Since it was first proposed by Charles F. Scott, it is frequently referred to as Scott connection. This connection can also be used for 3-phase to 2-phase transformation as explained.

One of the transformers has centre taps both on the primary and secondary ending (Fig1.14) and is known as the main transformer. It forms the horizontal member of the connection (Fig.1.15).

The other transformer has a 0.866 tap and is known as teaser transformer. One end of both the primary and secondary of the teaser transformer is joined to the centre taps on both primary and secondary of the main transformer respectively as shown in Fig. 1.15(a) . The other end A of the teaser primary and the two ends B and C of the main transformer primary are connected to the 3-phase supply.

The voltage diagram is shown in Fig 1.15(a) where the 3-pahse supply line voltage is assumed to be 100 V and a transformation ratio of unity. For understanding as to how 3-phase transformation results from this arrangement, it is desirable to think of the primary and secondary vector voltage forming geometrical Ts' (from which this connection gets its name).

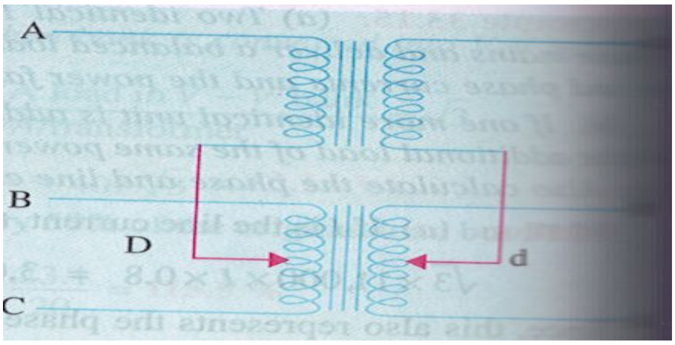


Fig1.14

In the primary voltage T of Fig 1.15(a) E_{DC} and E_{DB} are each 50V and differ in phase by 180^0 because both coils DB and DC are on the same magnetic circuit and are connected in opposition. Each side of the equilateral triangle represents 100 V. The voltage E_{DA} being the altitude of the equilateral triangle is equal to ($\sqrt{3}/2$) X100=86.6 V and lags behind the voltage across the main by 90^0. The same relation holds good in the secondary winding so that *abc* is a symmetrical 3-phase system.

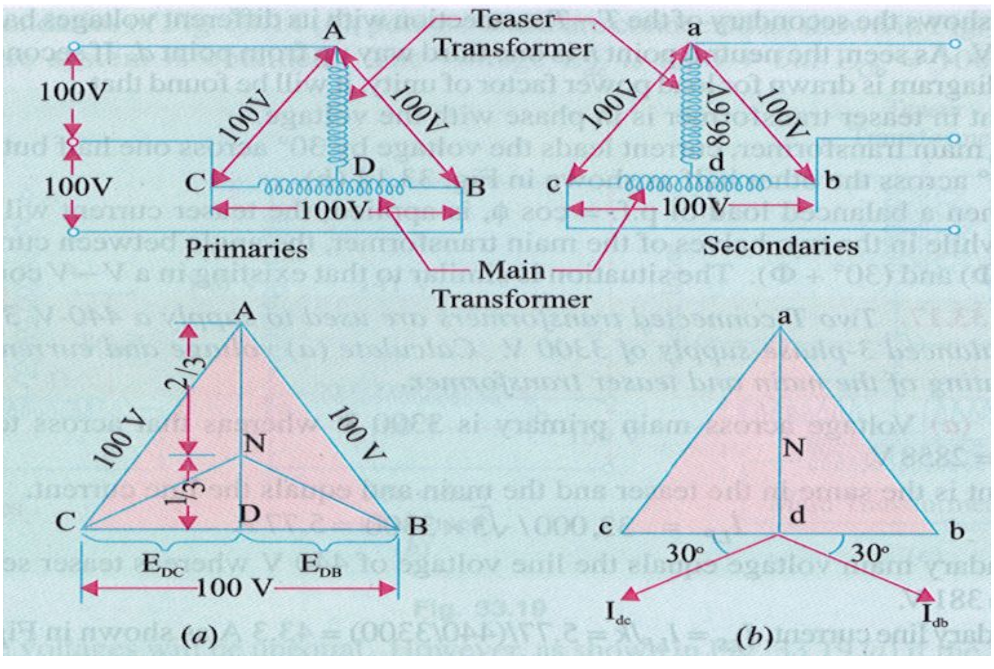


Fig.1.15

With reference to the secondary voltage triangle of Fig. 1.15(b),it is should be noted that for a load of unity power factor, current I_{db} lags behind voltage E_{db} by 30^0 and I_{db} leads

E_{dc} by 30^0 . In other words, the teaser transformer and each half of the main transformer, all operate at different power factors.

Obviously, the full rating of the transformers is not being utilized. The teaser transformer operates at only 0.866 of its rated voltage and the main transformer coils operate at $\cos 30^0 = 0.866$ power factor, which is equivalent to the main transformer's coils working at .86.6 per cent of their kVA rating. Hence the capacity to rating ratio in a T-T. Connection is 86.6%- the same as in V-V connection if two identical units are used, although heating in the two cases is not the same.

If, however, both the teaser primary and secondary windings are designed for 86.6volts only, then they will be operating at full rating, hence the combined rating of the arrangement would become (86.6+86.6)/(100+86.6)=0.928 of its total ratting. In other words, ratio of kVA utilized to that available would be 0.928 which makes this connection more economical than open-Δ with its ratio of 0.866.

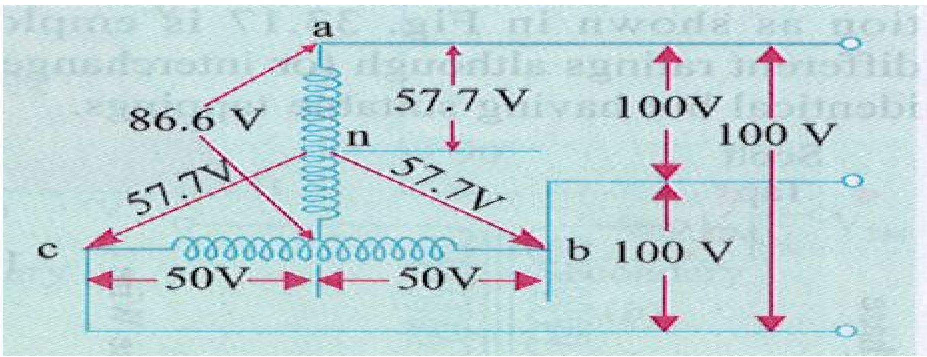


Fig1.16

Fig1.16 shows the secondary of the T-T connection with its different voltages based on a nominal voltage of 100 V. As seen, the neutral point n is one third ways up from point d. If secondary voltage and current vector diagram is drawn for load power factor of unity, it will be found that

1. Current in teaser transformer is in phase with the voltage.
2. In the main transformer, current leads the voltage by 30° across one half but lags the voltage by 30° across the other half as shown in figure 33.15(b)

Hence when a balanced load of power factor =COSΦ, is applied , the teaser current will lag or lead the voltage by Φ while in the two halves of the main transformer, the angle between current voltage will be (30°-Φ) and (30°+Φ). The situation is similar to that existing in a V-V connection.

Example-1.4. Two T- connected transformers are used to supply a 440V, 33KVA balanced load from a balanced three phase supply of 3300V. Calculate (a) Voltage and current rating of each coil (b) KVA rating of the main and teaser transformer.

Solution :- (a) Voltage across main primary is 3300V where as that across teaser primary is =0.866 X 3300 =2858V

The current is the same in the teaser and the main and equals the line current.

ILP=33000/√3X3300=5.77A

The secondary main voltage equals the line voltage of 440V whereas teaser secondary voltage =0.866 X 440=381V

The secondary line current, I**ls**= **ILP**/k=5.77/(440/3300)=43.3A as shown in figure1.17

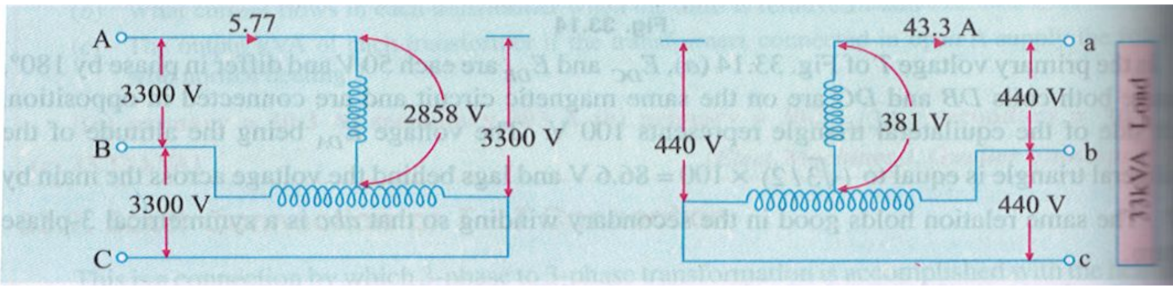


Fig.1.17

(b) Main KVA= 3300 X 5.77 X 10^{-3}=**19KVA**

Teaser KVA = 0.866 X main KVA= 0.866 X 19= **16.4KVA**

1.12. Three-phase to Two-phase Conversion and vice-versa

This conversion is required to supply two-phase furnaces, to link two-phase circuit with 3-phase system and also to supply a 3-phase apparatus from a 2-phase supply source. For this purpose, Scott connection as shown in fig 1.18 is employed. This connection requires two transformers of different ratings although for interchangeability and provision for spares, both transformers may be identical but having suitable tappings.

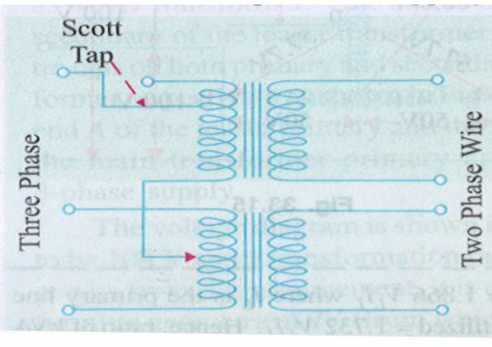


Fig1.18

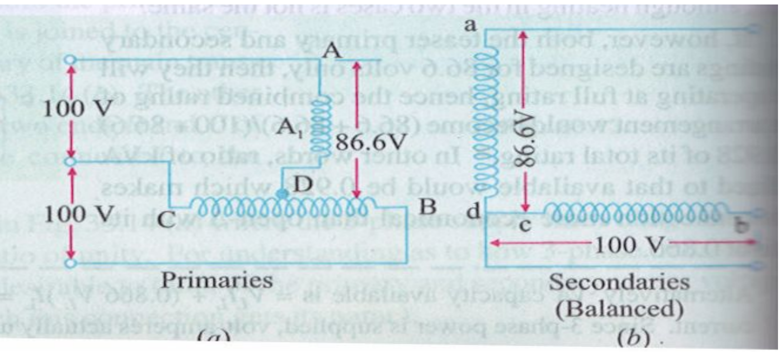


Fig.1.19

If, in the secondaries of Fig.1.15 (b), points c and d are connected as shown in Fig.1.19 (b), then a 2-phase, 3-wire system is obtained. The voltage E_{dc} is 86.6 V but E_{cb} = 100V, hence the resulting 2-phase voltages will be unequal. However, as shown in Fig.1.20 (a) if the3-phase line is connected to point A_1, such that DA_1 represents 86.6%of the teaser primary turns (which are the same as that of main primary), then this will increase the volts/turn in the ratio of 100:86.6, because now 86.6 volts are applied across 86.6 percent of turns and not 100% turns. In other words, this will make volts/turn the same both in primary of the teaser and that of the main transformers. If the secondaries are of both the transformers have the same number of turns, then the secondary voltage will be equal magnitude as shown, thus resulting in a 2-phase, 3-wire system.

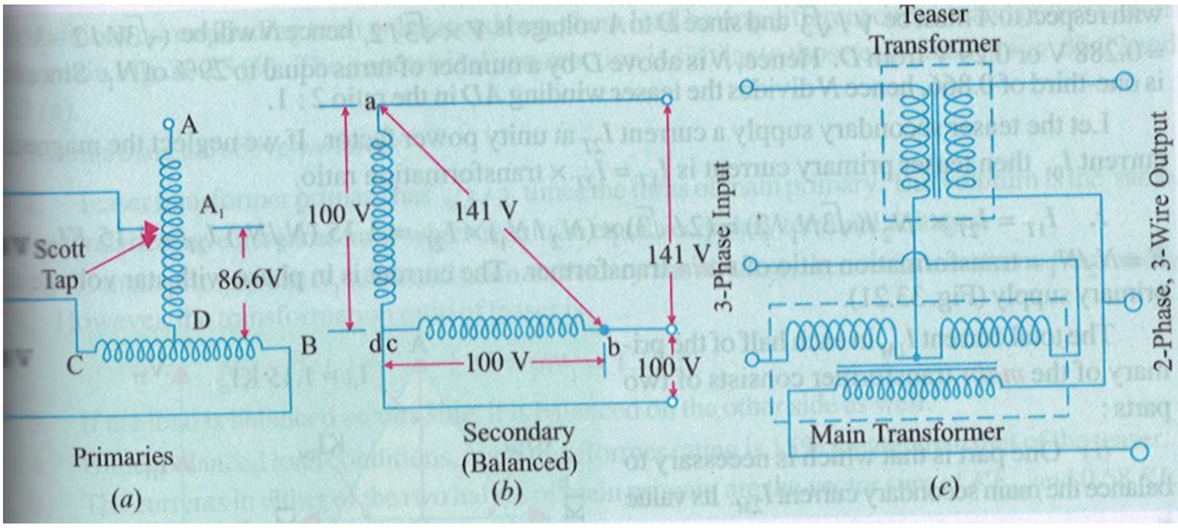
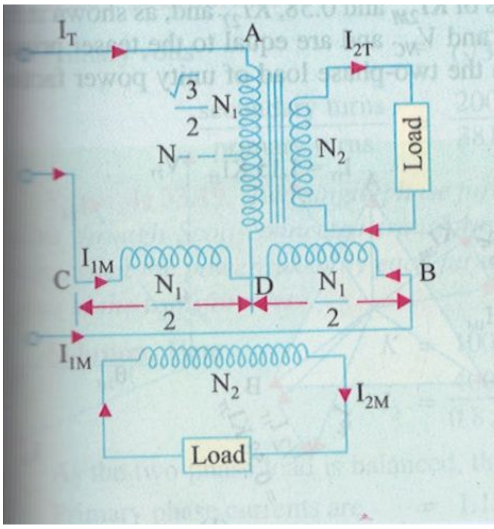


Fig1.20

Consider the same connection drawn slightly differently as in fig.1.21. The primary of the main transformer having N_1 turns is connected between terminals CB of a 3-phse supply. If supply line voltage is V. then obviously V_{AB}=V_{BC}=V_{CA}= V but voltage between A and D is V x $\sqrt{3}/2$. As said above, the number of turns between A and D should be $(\sqrt{3}/2)N_1$

for making volt/turn the same in both primaries. If so, then secondaries having equal turns , the secondary terminal voltages will be equal in magnitude although in phase quadrature.

-12-



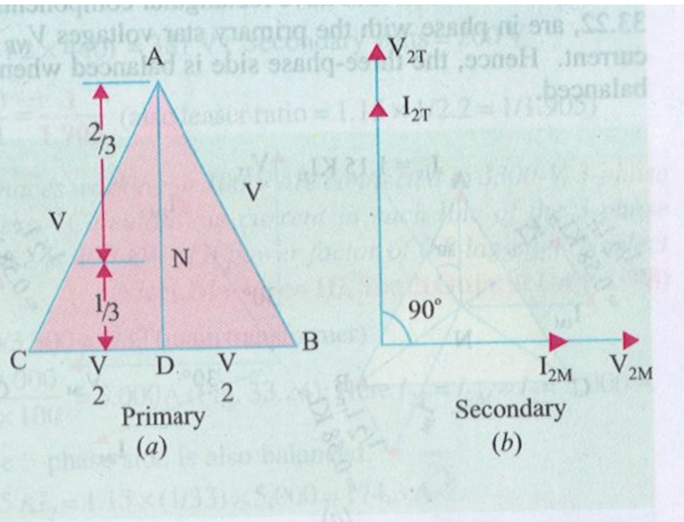


Fig1.21

Fig1.22

It is to be noted that point D is not the neutral point of the primary supply because its voltage with respect to any line is not $V/\sqrt{3}$. Let N be the neutral point. Its position can be determined as follows. Voltage N of with respect to A must be $V/\sqrt{3}$ and since D to A voltage is $V \times \sqrt{3}/2$, hence N will be $\sqrt{3}V/2 - V/\sqrt{3} = 0.288V$ or 0.29V from D. Hence, N is above D by a number of turns equal to 29% of N_1. Since 0.288 is one third of 0.866, hence N divides the teaser winding AD in the ratio 2:1.

Let the teaser secondary supply a current I_{2T} at unity power factor. If we neglect the magnetizing current I_o, then teaser primary current is $I_{1T} = I_{2T} \times$ transformation ratio

$\therefore$ $I_{1T}=I_{2T} \times N_2/(\sqrt{3}N_1/2) = (2/\sqrt{3}) \times (N_2/N_1) \times I_{2T} = 1.15 \times (N_2/N_1) \times I_{2T} = 1.15K\, I_{2T}$ where $=N_2/N_1$=transformation ratio of main transformer. The current is in phase with star voltage of the primary supply (figure1.22)

The total current I_{1M} **in each half of the of the primary of the main transformer consists of two parts:**

1. **One part is that which is necessary to balance** the main secondary current I_{2M}, its value is $= I_{2M} \times (N_2/N_1) = K\, I_{2M}$

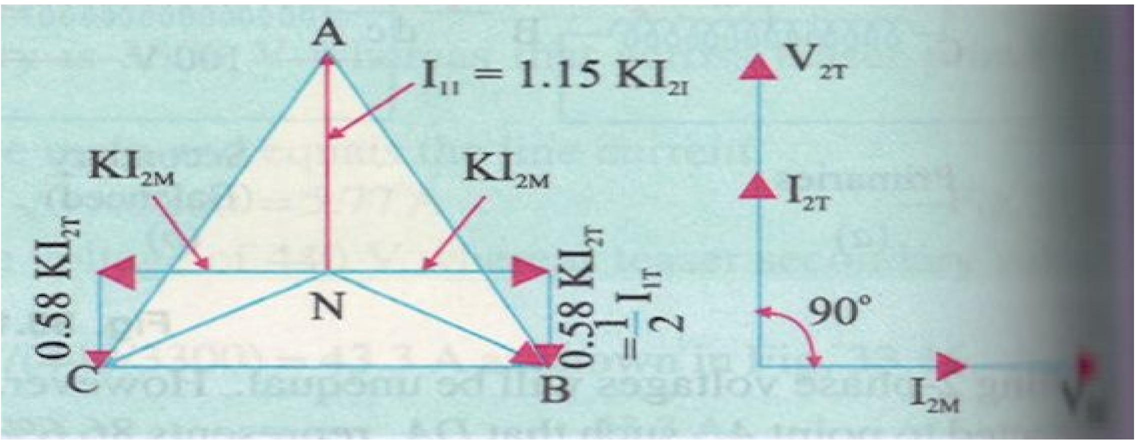


Fig1.23

2. The second part is equal to one half of the teaser primary current i.e. $0.5\ I_{1T}$. This is so because the main transformer primary forms a return path for the teaser primary current which divides itself into two halves at mid point D in either direction. The value of each half is $=0.5\ I_{1T}=1.15K\ I_{2T}/2= 0.58\ K\ I_{2T}$.

Hence the current in the lines B and C are obtained vectorially as shown in fig.1.23. It should be noted that as the two halves of the teaser primary current flow in opposite directions from point D, they have no magnetic effect on the core and play no part at all in balancing the secondary ampere-turns of the main transformer.

The line currents thus have rectangular components of $K\ I_{2M}$ and $0.58\ K\ I_{2T}$ and as shown in fig. 1.23, are in phase with primary star voltages V_{NB} and V_{NC} and are equal to the teaser primary current. Hence, the three phase side is balanced when the two phase load of unity power factor is balanced.

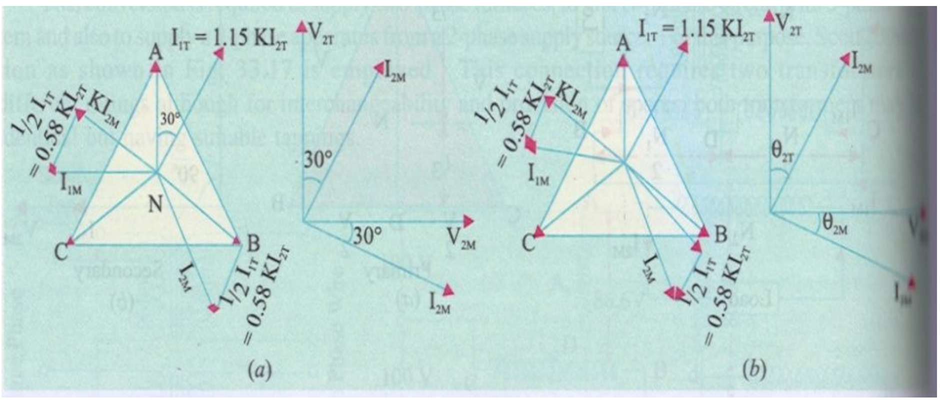


Fig1.24

Figure 1.24(a) illustrates the condition corresponding to a balanced two phase load at a lagging power factor of 0.866. The construction is the same as in figure 1.23. it will be seen that the three- phase side is again balanced. But under these conditions the main transformer rating is 15%greater than that of the teaser, because its voltage is 15% greater although its current is the same.

Hence, ***we conclude that if the load is balanced on one side, it would always be balanced on the other side.***

The conditions corresponding to an unbalanced two-phase load having different currents and power factors are shown in figure 1.24(b). The geometrical construction is similar to those explained in figure 1.23 and 1.24(a).

Summarizing the above we have:

1. Teaser transformer primary has $\sqrt{3}/2$ times the turns of main primary. But volt/turn is the same. Their secondaries have the same turns which result in equal secondary terminal voltages.

2. If main primary has N_1 turns and main secondary has N_2 turns, then main transformation ratio is N_2/N_1. However, the transformation ratio of teaser is $N_2/(\sqrt{3}\,N_1/2) = 1.15\,N_2/N_1 = 1.15K$.

3. If the load is balanced on one side, it is balanced on the other side as well.

4. Under balanced load conditions, main transformer rating is 15% greater than that of the teaser.

5. The currents in either of the two halves of main primary are the vector sum of $K\,I_{2M}$ and $0.58K\,I_{2T}$(or $0.5\,I_{1T}$)

Example 1.5. Two transformers are required for Scott connection operating from a 440V, three phase supply for supplying two single phase furnaces at 200V on the two phase side. If the total output is 150KVA, calculate the secondary to primary turn ratio and the winding currents of each transformer.

Soln: - **main transformer**

Primary volts= 440V ∴ secondary volts=200V ∴ N_2/N_1=200/440=1/2.2

Secondary current = 150000/2 X 200=**375 A**

Primary current =375 X 1/2.2=**197A**

Teaser transformer

Primary volts= (√3/2 X 440) =381V: Secondary volts =200V

Secondary turns /primary turns =200/381=**0.52**(also teaser ratio =1.15 X 1/2.2=0.52).

1.13. Parallel operation of three phase transformer

Transformers are said to be connected in parallel when their primary windings are connected to a common voltage supplier and their secondary windings are connected to a common load.

1.14. Reasons for parallel operation

1. Extension of loads - for large loads it may be impracticable or uneconomical to have a single large transformer.

2. Capacity to spare – in substations the total load required may be supplied by an appropriate no of transformers of standard size. This reduces the spare capacity of the substation.

3. Future extension - there scope of future extension of a substation to supply a load beyond the capacity of the transformers already installed.

4. If there is a breakdown of transformer in system of transformers connected in parallel, there is no interruption of power supply for essential service. Similarly when a transformer is taken out of service for its maintenance and inspection the continuity of supply is maintained.

1.15. Condition for parallel operation

All the condition which applied to the parallel operation of single phase transformer also is applied to the parallel running of three phase transformer but with the following addition

1. The voltage ratio must refer to the terminal ***voltage of primary and secondary.*** It is obvious that this ratio may not be equal to the ratio of the number of turns per phase. For example, if V1, V2 are the primary and secondary terminal voltages, then for Y/Δ connection the turn ratio is $V2/ (V1/\sqrt{3}) = \sqrt{3}V2/ V1$.

2. The phase displacement between primary and secondary voltages must be the same for all transformers which are to be connected for parallel operation.

3. Phase sequence must be the same.

4. All the three transformers in the three phase transformer bank will be of the same construction either core or shell.

Note 1: IN dealing with three phase transformer calculation are made for one phase only. The value of equivalent impedance used is the impedance per phase referred to secondary.

2. In case the impedance of primary and secondary windings are given separately then primary impedance must be referred to secondary by multiplying it with (transformation ratio)^2.

3. For Y/Δ or Δ/ Y transformers should be remembered that the voltage ratio as given in the question is referred to terminal voltages and are quite different from turn ratio.

1.16. Tap changers in transformers

The modern equipments, utilising electrical energy are design to operate satisfactorily at one voltage level. It is therefore of paramount importance to keep the consumers' terminal voltage, within the prescribed limits. The transformer output voltage and hence the consumers' terminal voltage, can be controlled by providing taps either on the primary or on the secondary.

The principle of regulating the secondary output voltage is based on changing the number of turns in the secondary quantities. $V_2 = (N_2/ N_1) \times V_1$.

If the tap changer is design to operate, when the transformer is out of circuit, it is then called off-load (or no load) tap changer. A tap changer design to operate with the transformer in the circuit is called on load tap changer.

1.17. No load (or off-load) Tap changer

This tap changer is used for seasonal voltage variation. And elementary form of no load tap changer is illustrated in figure.1.25. It has six studs mark from one to six. The winding is tapped at six points equal to the number of studs. The tapping leads are connected to six correspondingly marks stationary studs arranged in circle. The face plate carrying the six studs, can be mounted anywhere on the transformer, say on the yoke or any other convenient place. The rotatable arm R can be rotated by means of hand wheel, from outside the tank.

If the winding is tapped at 2.5% intervals, than with the rotatable arm R

- At studs 1,2: Full winding is in circuit
- At studs 2,3: 97.5% of the winding is in circuit
- At studs 3,4: 95% of the winding is in circuit
- At studs 4,5: 92.5% of the winding is in circuit
- At studs 5,6: 90% of the winding is in circuit

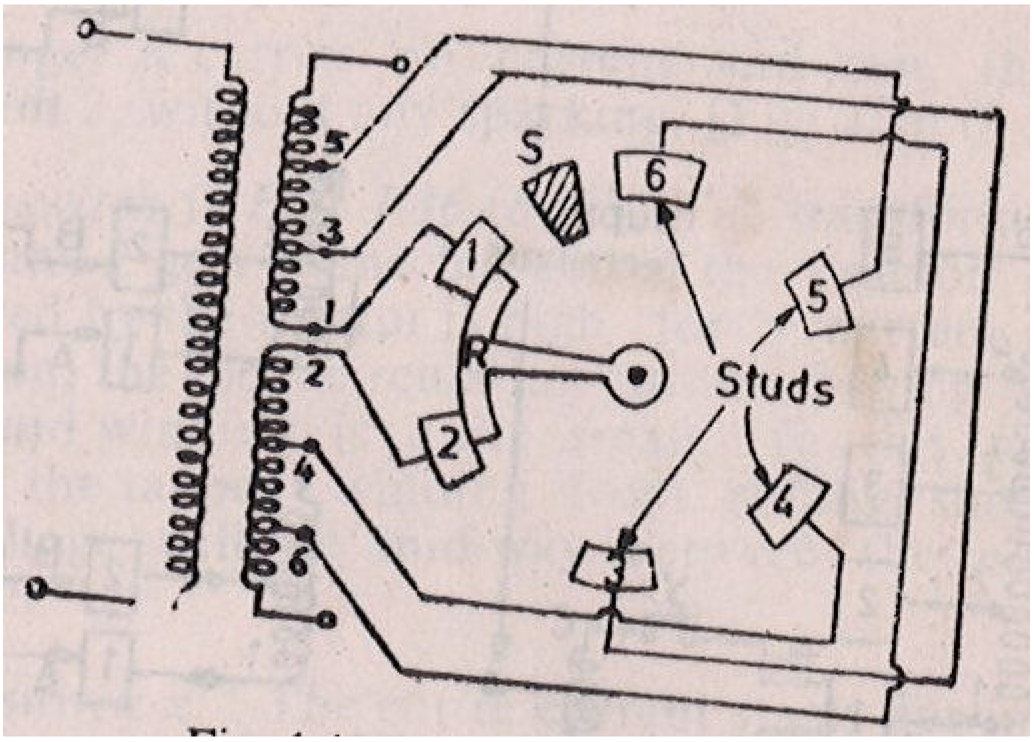


Figure1.25

Stop S fixes the final position and prevents the arm R from being rotated clockwise. In the absence of stop S, the arm R may come in contact with studs 1 and 6. In such a case, only the lower part of the winding is cut out circuit and this is undesirable from mechanical stress considerations.

The tap changing must be carried out only after the transformer is disconnected from the supply. Suppose arm R is at studs 1, 2. For bringing arm R at studs 2and 3, the transformer is first de-energised and then arm R is rotated two bridge studs 2 and 3. After this, transformer is switched on to the supply and now 97.5% of the winding remains in circuit.

1.18. On-load tap changer

The tap changer s used for daily or sort period voltage alteration. The output voltage can be regulated with the changer, without any supply interruptions. During the operation of an on-load tap changer.

- The main circuit should not be opened otherwise dangerous sparking will occur and
- No part of the tap winding should get short circuited

One form of elementary on load tap-changer is illustrated in figure 1.26(a). The centre tap reactor C prevents the tapped from getting short circuited. The transformer tapings are connected to the correspondingly marked segment s 1 to 5. Two moveable fingers, A and B connected to centre-tapped reactor via. Switches x and y, make contact with anyone of the segments under normal operations.

In fig.1.26 (a), both the fingers are in contact with segment 1 and full winding is in circuit switches x, y are closed. One half of the total current flows through x, lower half of the reactor and then to the external circuit. It is seen that currents in the upper and lower halves of the reactor flow in opposite direction. Since the whole reactor is wound in the same direction the m.m.f produced by one half is opposite to the m.m.f produced by the secondary half. These m.m.fs are equal and the net m.m.f is practically zero: therefore the reactor is almost non inductive and the impedance offered by it is very small. Consequently the voltage drop in the centre-tap reactor is negligible.

When a change in voltage is required the finger A and B can be brought to segment to, by adopting the following sequence of operations.

- Open switch y figure 1.26 (b1). The entire current must now flow through the lower half of the reactor. It therefore, becomes highly inductive and there is a large voltage drop. It should be noted that the rector must be designed to handle full load current, momentarily.
- The finger B carries no current and can therefore, be moved to segment 2, without any sparking (figure 1.26(b2)).
- Close switch Y figure1.26 (b3) the transformer winding between taps 1 and 2 gets connected across the reactor. Since the impedance offered by the reactor is high for a current flowing in only one direction, the local circulating current flowing through the reactor and tapped winding is quite small. In this manner, the reactor prevents the tapped winding from getting short circuited. The terminal voltage will be mid-way between the potentials of tappings 1 and 2.
- Open switch x: The entire currents start flowing through the upper half of the reactor, manifested by large voltage drop, fig.1.26 (b4).
- Move the finger A from segment 1 to segment 2 and then close switch x: The winding between taps 1 and 2 is therefore completely out of circuit, fig. 1.26(b5). If further change in voltage is required, the above sequence of operations is repeated.
- For large power transformers the switches x and y may be circuit breakers.

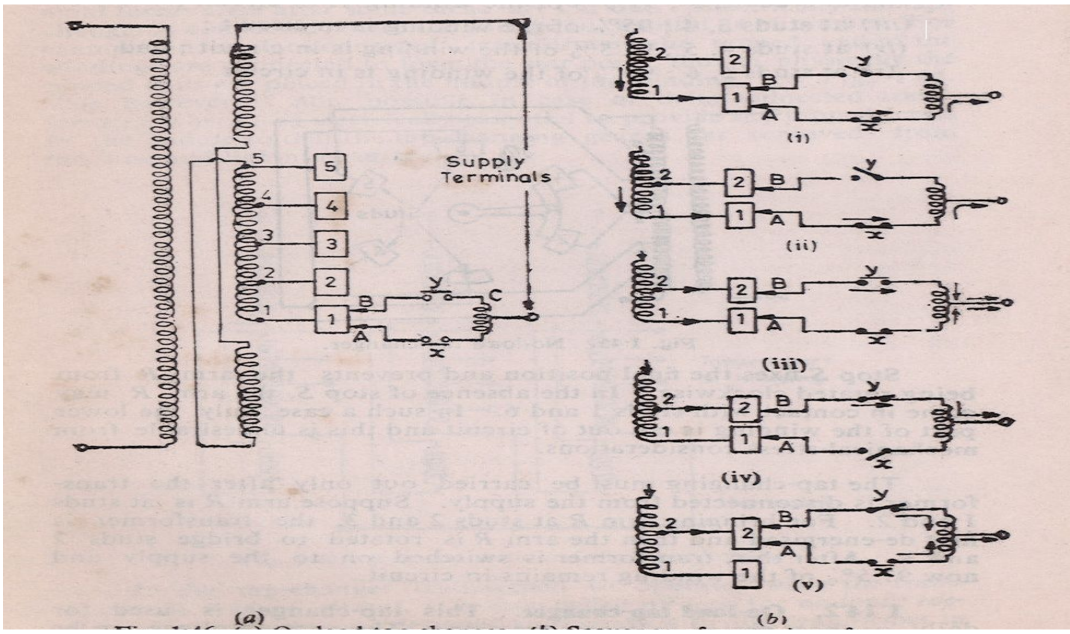


Figure1.26

1.19. Maintenance of Transformer

The normal life of transformer is about 30 years. It could even be longer if operated carefully and maintained regularly. The main object of maintenance of transformers is to maintain its insulation in good condition. Factors affecting the insulation of a transformer are: moisture, presence of oxygen, and solid impurities.

Maintenance of transformers needs (i) external inspection, and (ii) internal inspection periodically.

The external inspection requires inspection of parts and auxiliaries of the transformer that can be done without opening the tank or lowering the oil level but with the transformer taken out of service,e.g. megger tests, ratio tests, water –flow tests, taking out sample of oil and testing it, inspection of bushings,breathers,oil level,tank,gaskets,ground wire for all auxiliary apparatus,etc.In the case of large transformers, the condition of circulating pumps, on load tapping gears, oil gauges, pressure relief devices, oil gauges,etc.,need to be checked.

1.20. Checking and testing of transformer oil

The deterioration of insulation oil is generally due to oxidation, especially when the transformer works under the condition of high temperatures. Oxidation is due to the formation of acids, sludge and water which accompanies the chemical change. Samples of transformer oil are taken out carefully and tested for colour and odour. Cloudiness in oil may be due to suspended moisture or suspended solid matter. Dark brown colour may indicate dissolved asphalt, green colour dissolved copper compounds; and acid smell indicates the presence of volatile acids.

The oil samples may be tested as follows:

(a) The dielectric strength of oil should be tested as per IS: 335 – 1953.The oil should withstand the test voltage of at least 30kV for one minute without breakdown.

(b) Cracle test for free water should be performed as per IS: 335 – 1953.The test is only qualitative.

(c) The acidity of oil should be determined as per IS: 1866 – 1961.

(d) Sludge test: The traces of solid matter in oil samples may be examined as per IS: 1866 – 1961.

1.21. Insulation resistance

The insulation resistance is measured by megger test along with the temperature. This is because the insulation resistance in megohms gets reduced to nearly half for every 10^0c temperature rise. The insulation resistance should not be less than two megohms for each 1000 v of operating voltage.

1.22. Internal inspection of Transformer

Take samples of oil from top and bottom for testing: Lower oil in transformer. Check inside bushings, brackets, HV, LV windings for damage insulation; check connections, ground of core, insulation condition of various parts, and inside condition. An inspection schedule should be drawn for checks monthly, quarterly and yearly inspection.

1.23. Maintenance schedule

Every hour: Check temperature of oil, windings, ambient, load & voltage. Adjust load to keep the temperature rise within a permissible limit.

Daily: (a) Check oil level; if low, fill in dry oil.

(b) Check the colour of the silica gel in the breather. Colour should be blue. If the colour of the silica gel becomes pink replace them.

Quarterly: Check for proper working of cooling fans, circulating pumps, etc.

Half-yearly: Check the dielectric strength of oil, bushes, insulators, cable boxes, filter, and replace oil if necessary.

Yearly: Check oil for acidity, sludge formation, contacts, lightning arrestors, etc.

Check alarms, relays, etc.

Check earth resistance.

Five yearly: Carry out overall inspection of the transformer including lifting of core and coils. Clean the transformer with dry transformer oil.

1.24 Diagnostic tests for power Transformer

The Transformer is the heart of the Grid S/S & it is the most costly equipment in it. Any failure in it will not only damage the equipment nearby & may also create danger to the life of the Operating staff. It takes a lot of time to replace a power Transformer, which will affect the steady power supply. Hence it becomes very essential to ascertain the condition of the Transformer under service. The monitoring of Transformer's condition is not that simple as it sounds. Because no test give a very clear picture about the condition. So to ascertain the real condition of the Transformer diagnostic analysis has to be done from a set of results. This is known as diagnostic analysis of a power Transformer.

Normally every utility make some routine tests at least annualy to the transformer. Any slight deviation in the routine test, diagnostic analysis may refer to. Even this analysis is essential at a new condition for signature impression as well as to detect any design or assembly defect. The total life of the Transformer may be devided in three segmants. The initial period, which is a small period usually 4 to 5 years, is known as **INFANT MORTALITY.** The percentage of failure is quite high in this period. Any failure in this period attributes to design or assemble failure. The 2nd stage which is quite longer period is the **NOMAL PERIOD** & has a very less percentage of failure. In this period attribute to poor maintenance. The 3rd period is the **AGEING PERIOD.** Again the persentage of failure increases in this period bacause of ageing factor of mainly soild insulation used. The Frequency of analysis should be more may be almost in every 2 years. This life cycle characteristic is known as **BATH TUB CHARECTETISTIC** because of shape.

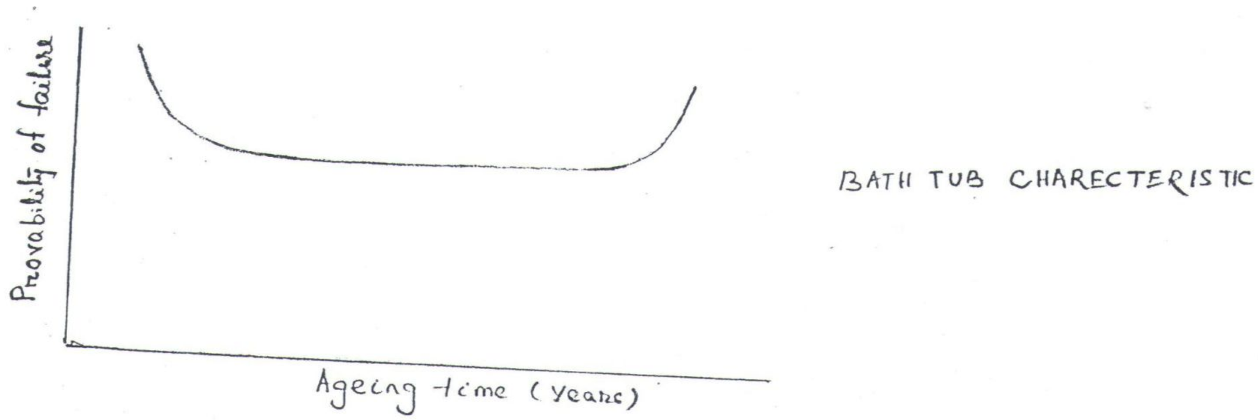


Fig.1.27

The normal routine test that may be conducted atleast every year are;

- IR value.
- PI value.
- Trans value
- BDV Test value.
- However, BDV may be done aditionaly atleast in every 2 months.
- If any abnormalities are found in the above tests, then only we have to go for dissolved gas analysis (DGA Test), otherwise it is not required.

The life of the Transformer is generally the life of the soild insulation, the cellulosic paper. The degree of deterioration of the insulation is mainly due to the diffrent stress that act on the transformwer under service, which reflects on the life of the transformer used, The stree that act.

- **Mechanical stress:** between conductors, leads & windings due to over currents or fault current, mainly due to system short circuits.
- **Thermal Stress:** Due to local heating, over load currents & leakage flux or due to malfunctioning of cooling system.
- **Dialectric Stree:** System over voltage, Transient impluse condition or due to internal resonance within a winding.
- **Environmental Stress:** Moisture ingress, pollution.

Even may new type of insulation has been developed, the cellulosic paper is still widely used.

The rising temperature in presence of moisture & oxygen accelerate the aging process of the soild insulation, For a example, the paper with 2% moisture ages three times taster then 1% moisture & 30 times to 3% moisture content. The degradation product from oil oxidation, such as peroxides & water soluble acid absorbs in paper & makes it brittle & low strength oxycellulose. The oxidation gradully deplets the natural oxidation inhibitors present in naphthalenic oil & products are acid, ketones, peroxides, soap, and aldehyds. This causes colloidal contamination in the oil which form hydrocorbon which again polymerisis to form partly conduting sludge & get deposited on the windings thus it makes heat transfer more diffcult & oxidation become more faster due to rise of temperture, So it is conclusive that presence of moisture & oxygen in oil or paper is the main culprit to reduce life of the transformer. The routine test must be conducted regularly to know the presence of the moisture & weather it is within the limit or not. If the value is low then there is no problem othewise we have to go for further analysis regrading the presence of moisture & other conducting gaese & where it is present (whether in oil or in paper or in both).A accordingly steps will be taken. The oxidations also accelerate due to partial discharge.

By now our stand is more clear that;

- We want to know whether any moisture or any conducting soluble gas or conducting particulars present in the insulation.
- If present not within limit then it is essential to know where it is and in which formand how to separate it out & to increase the life period.
- We should not allow to increase the moisture content in oil and if however it has entered then it is assentail to knwo to what level the damage has been taken place. So that we can decrease the effect to certain level and increase the life of our transformer.

1.25. Routine test

(1) IR value:

It is simlpy the insulation resistance of the insulating materials i.e. paper & oil in combination, A DC petential is applied usally 5 KV between diffrent windings, between winding & tank of the transformer. Earlier, the value was noted after allowing the current for 15 sec. But now a day value is noted after 1 min. As the real values can be known only after allowing the current for certain time. What should be the IR Value? It is a real debate. It depends upon the size & shape of the insulating materials & also affected by different environmental condition. In a thumb rule people consider it as 1.5 MΩ/ KV. If there is any huge variation then it is generally marked, Before taking the IR value all clamps & connectors should be propely tightened & bushing & tanks should be cleaned. This test has least importance unless & untill the value is out right low.

(2) PI Value:

It is known as polarisation index. It is a number having no uint. It is a ratio of insulation resitence value taken for 10 min. to 1 min. Now the question arises what is polarisation & how its value is affected due to the presence of moisture or any conducting soluble gas.

In a conductor there is free electron, which is free to move under application of external field, but in case of insulator there is no free electron. At normal condition the electron moves arround the protons such that CG of both consider with each other, has no net polar effect, when an external field is applied the rotation of the electron arround the proton is no more circular but eccentric as shown in the figure.

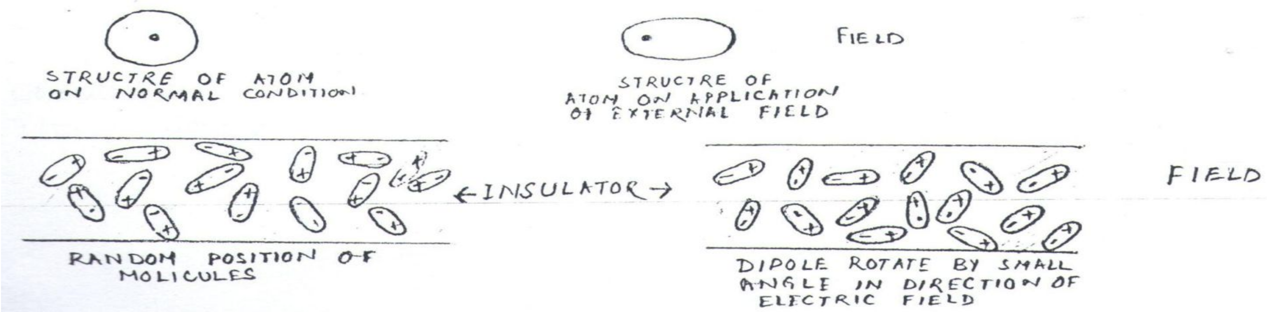


Fig.1.28

This implies when an electrical field is applied the CG of proton & electron are no more same but displaced with a small gap. This result into a electrical dipole or it can be said that polarisation has taken place. This dipoles orient around itself in such a manner that the net electrical field produced by the dipoles, opposes the applied electrical field. The reduction in applied field reduces the current or increase the resistance values. The increasing value is more & more as more & more dipole oriented around itself. After around 10min almost all orientation takes place so 10 min values is taken.T he polarisation index has a other name that is **DRYNESS FACTOR**.

The name suggests that the dryness of the insulation has a certain role over that ratio which is known as polarisation index. If there is some moisture or desolve conducting gases present in a insulator then a conduction sphere appears around the insulator which does not allow to penetrate the external field.This reduces the polarisation effect.So reductuion & PI value indicates the presence of moisture or any desolve conducting gases in the insulator,as per IS the value above 1.5 is consider to be good.

3) Tan δ value:

The PI value is affectec by moisture & desolve gases but there may be many other conducting non soluble substances which allowed more current to flow to the insturator causing more heat & oxidation.Thus causing detoriation of insulating materials.Tan δ test gives more clear-cut picture regarding the presence of any conducting materials presence in the insulator.

When a insulator is in between two conducting substances it is nothing but a capacitor.So when we apply a AC potential between two winding or winding & tank which is earthed acts as a capacitive circuit as both solid & liquid insulator are in between .Idealy the current should lead the voltage by an angle of 90^0 .But practically it will not beacuase of certain resistance present in it.The angle by which it falls to reach 90^0 is known as δ angle.

Fig.1.29

$\text{Tan}\delta = V_R / V_C = R / X_C$

Higher is the value of Tanδ, more is the resistive materials present in the insulation in any form as per value upto 0.2 allowed.

To major the Tanδ value the instrument used is nothing but a Sheraing bridge, supported by a software to give the result directly in Tanδ.

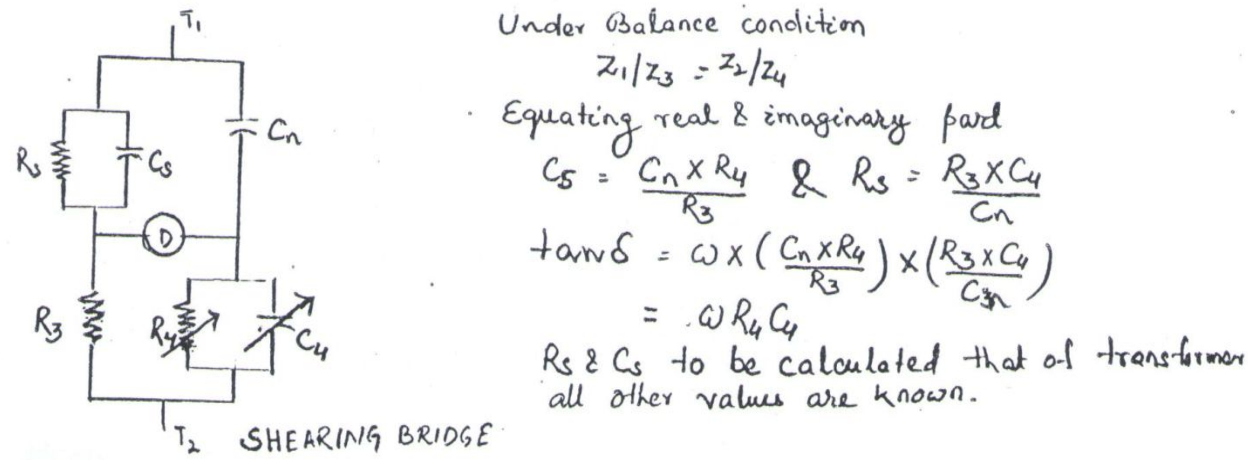


Fig.1.30

4) BDV Test value:

It is a very simple test. The breakdown voltage (BDV) of an insulator is the potential at which it loses its insulating property & become conducting. Oil is taken in a glass or plastic container of usually 300ml to 500ml capacities. The electrode are of copper, brass bronze or stainless steel well polished having spherical shape of dia 12.5mm to 13mm separated by 2.5± 0.1mm.

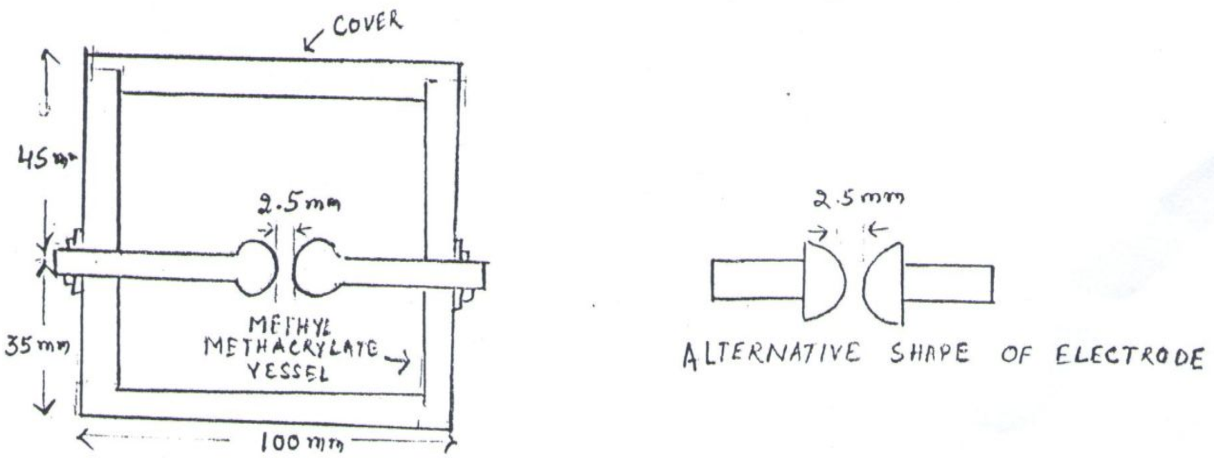


Fig.1.31

The oil under test should be between 15^0C to 35^0C preferably 27^0C.The applied potential at rated frequency should be raised gradually at a rate around 2KV per sec till flash over takes place. The test kits automatically switch OFF within 0.02sec, The average of six tests result is taken. The time interval between two tests should be 5min. if the disappearance of air bubble does not take place. The value recommended by IS is above 50KV.This test may be taken in every month. Proper care should be taken at sampling time; so that no external moisture enters to it.

SUMMARY

BDV	>50KV <50KV	Good. Should be taken again in a better weather condition & if it is still low, then filtration or dehydration may be required & will be decided after other tests.
PI	>1.5 <1.5	Good. Filtration or dehydration may be required & will be decided after other tests.
Tanδ	<0.2 >0.2	Good. Filtration or dehydration may be required & will be decided after other tests.

Other Test:

1) IR, PI & Tanδ.
2) Test on Oil & DGA.
3) Recovery Voltage Measurement.
4) Dielectric Spectroscopy Test.
5) Magnetic Balance Test.
6) Turns Ratio Test.
7) Frequency Response Analysis.
8) Coil Resistance Test.
9) Degree of polymerisation Test. (DP)
10) Partial Discharge Test (PD)
11) Surge Voltage Analysis Test.

GANDHI ACADEMY OF Technology & Engineering
GOLANTHARA, BERHAMPUR

Lecture Notes on

Power Electronics and PLC

For 5th Semester

Electrical Engineering (Diploma)

As per SCTE&VT Syllabus

PREPARED BY: SOUMYA SHYMALAI MAHAPATRA

TH.5 POWER ELECTRONICS AND PLC

Name of the Course: Diploma in Electrical Engineering			
Course code:	Th.5	**Semester:**	5 th
Total Period:	60 Periods	**Examination:**	3 Hrs
Theory periods:	4 P / Week	**Internal Assessment:**	20
Tutorial:	-	**End Semester Examination:**	80
Maximum marks:	100		

A. Rationale:

The development of high power semiconductor devices has facilitated electronic control techniques for electrical power control in a simple, economic and efficient manner. Thus a new area of power electronics has now emerged which replaced the old and bulky method of power control through the use of small electronic devices. Power electronics application has occupied an indispensible position in industrial applications like heating, welding, uninterrupted power supply, battery charging etc. Industrial drives, lighting control are most efficiently controlled by power electronics devices to achieve optimum performance. The objective of this paper is to familiar students with the principles and operations of Power electronics devices in Industrial applications with drives control.

B. Objectives:

After completion of this subject the student will be able to:

1. Understand construction, working principle & application of various power electronics devices.
2. Know different gate triggering circuits and commutation methods.
3. Understand working principle of phase controlled rectifier.
4. Know the types and working principle of inverter.
5. Understand working principle and voltage control of chopper.
6. Understand frequency variation using Cyclo-converter.
7. Understand control principle of AC & DC industrial drive.
8. Know different application of SCR / Thyristor.
9. Concept in PLC & its Programming

C. TOPIC WISE DISTRIBUTION OF PERIODS

Sl. No	Topics	Periods
1	Understand The Construction And Working Of Power Electronic Devices	18
2	Understand The Working Of Converters, Ac Regulators And Choppers.	12
3	Understand The Inverters And Cyclo-Converters	08
4	Understand Applications Of Power Electronic Circuits	10
5	PLC And Its Applications	12
	Total	60

COURSE CONTENT:

1. UNDERSTAND THE CONSTRUCTION AND WORKING OF POWER ELECTRONIC DEVICES

Construction, Operation, V-I characteristics & application of power diode, SCR, DIAC,TRIAC, Power MOSFET, GTO &IGBT

Two transistor analogy of SCR.

Gate characteristics of SCR.

Switching characteristic of SCR during turn on and turn off.

Turn on methods of SCR.

Turn off methods of SCR (Line commutation and Forced commutation)

Load Commutation

Resonant pulse commutation

Voltage and Current ratings of SCR.

Protection of SCR

Over voltage protection

Over current protection

Gate protection

Firing Circuits

General layout diagram of firing circuit

R firing circuits

R-C firing circuit

UJT pulse trigger circuit

Synchronous triggering (Ramp Triggering)

Design of Snubber Circuits

2. UNDERSTAND THE WORKING OF CONVERTERS, AC REGULATORS AND CHOPPERS.

Controlled rectifiers Techniques(Phase Angle, Extinction Angle control), Single quadrantsemi converter, two quadrant full converter and dual Converter

Working of single-phase half wave controlled converter with Resistive and R-L loads.

Understand need of freewheeling diode.

Working of single phase fully controlled converter with resistive and R- L loads.

Working of three-phase half wave controlled converter with Resistive load

Working of three phase fully controlled converter with resistive load.

Working of single phase AC regulator.

Working principle of step up & step down chopper.

Control modes of chopper

Operation of chopper in all four quadrants.

3. UNDERSTAND THE INVERTERS AND CYCLO-CONVERTERS

Classify inverters.

Explain the working of series inverter.

Explain the working of parallel inverter
Explain the working of single-phase bridge inverter
Explain the basic principle of Cyclo-converter.
Explain the working of single-phase step up & step down Cyclo-converter.
Applications of Cyclo-converter.

4. UNDERSTAND APPLICATIONS OF POWER ELECTRONIC CIRCUITS

List applications of power electronic circuits.
List the factors affecting the speed of DC Motors.
Speed control for DC Shunt motor using converter.
Speed control for DC Shunt motor using chopper.
List the factors affecting speed of the AC Motors.
Speed control of Induction Motor by using AC voltage regulator.
Speed control of induction motor by using converters and inverters (V/F control).
Working of UPS with block diagram.
Battery charger circuit using SCR with the help of a diagram.
Basic Switched mode power supply (SMPS) - explain its working & applications

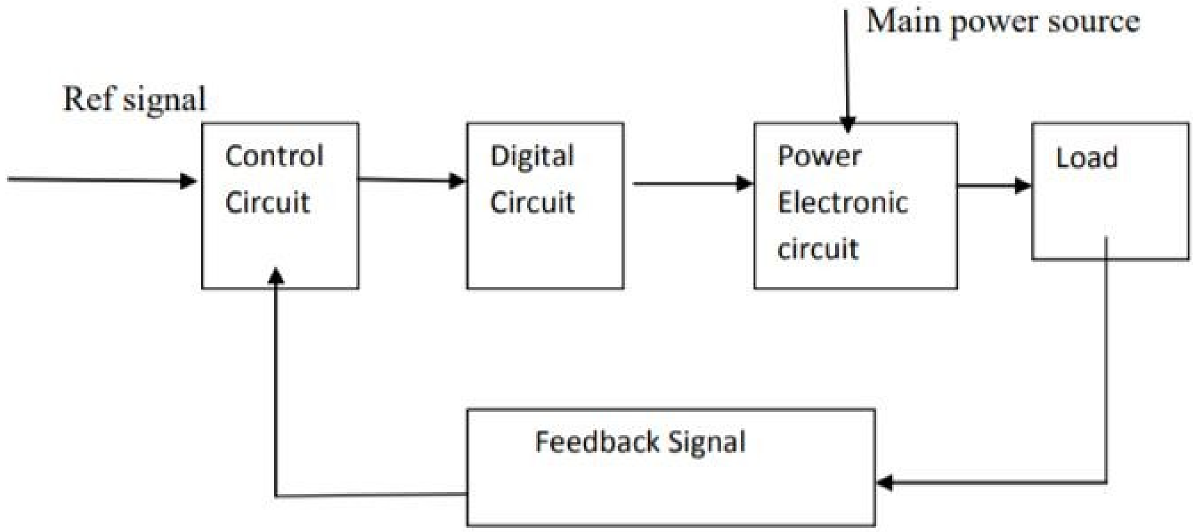
5. PLC AND ITS APPLICATIONS

Introduction of Programmable Logic Controller (PLC)
Advantages of PLC
Different parts of PLC by drawing the Block diagram and purpose of each part of PLC.
Applications of PLC
Ladder diagram
Description of contacts and coils in the following states
i)Normally open
ii) Normally closed
iii) Energized output
iv) latched Output
v) Branching
Ladder diagrams for i) AND gate ii) OR gate and iii) NOT gate.
Ladder diagrams for combination circuits using NAND, NOR, AND, OR and NOT
Timers-i)T ON ii) T OFF and iii)Retentive timer
Counters-CTU, CTD
Ladder diagrams using Timers and counters
PLC Instruction set
Ladder diagrams for following (i) DOL starter and STAR-DELTA starter (ii) Stair caselighting (iii) Traffic light Control (iv) Temperature Controller
Special control systems- Basics DCS & SCADA systems
Computer Control–Data Acquisition, Direct Digital Control System (Basics only)

POWER ELECTRONICS

INTRODUCTION

Power electronics may be defined as the application of solid state electronics for the control and conversion of electrical power. The inter relationship of power electronics with power, electronics and control is shown in the fig



Power electronics based on the switching of power semiconductor devices. With the development of power semiconductor technology, the power handling capabilities and switching speed of power devices have been improved tremendously.

Advantages of power electronics

- High efficiency due to low loss in power semiconductor devices.
- High reliability of power electronic converter system.
- Long life and less maintenance due to absence of any moving parts.
- Flexibility in operation
- Fast dynamic response compared to electromechanical converter system.
- Small size and less weight, thus low installation cost.

Application of power electronics

- Our Daily Life: fan regulator, light dimmer, air-conditioning, induction cooking, emergency lights, personal computers, vacuum cleaners, UPS (uninterrupted power system), battery charges, etc.
- Automotives and Traction: Subways, hybrid electric vehicles, trolley, fork-lifts, and many more. A modern car itself has so many components where power electronic is used such as ignition switch, windshield wiper control, adaptive front lighting, interior

lighting, electric power steering and so on. Besides power electronics are extensively used in modern traction systems and ships.

- Industries: Almost all the motors employed in the industries are controlled by power electronic drives, for eg. Rolling mills, textile mills, cement mills, compressors, pumps, fans, blowers, elevators, rotary kilns etc. Other applications include welding, arc furnace, cranes, heating applications, emergency power systems, construction machinery, excavators etc.
- Defense and Aerospace: Power supplies in aircraft, satellites, space shuttles, advance control in missiles, unmanned vehicles and other defense equipments.
- Renewable Energy: Generation systems such as solar, wind etc. needs power conditioning systems, storage systems and conversion systems in order to become usable. For example solar cells generate DC power and for general application we need AC power and hence power electronic converter is used.
- Utility System: HVDC transmission, VAR compensation (SVC), static circuit breakers, generator excitation systems, FACTS, smart grids, etc.

CONSTRUCTION, OPERATION, V-I CHARACTERISTICS & APPLICATION OF POWER DIODE, SCR, DIAC, TRIAC, POWER MOSFET, GTO &IGBT

POWER DIODE

Definition: A diode that has two terminals like anode & cathode and two layers like P & N, used in the power electronics circuits is known as power diode. This diode is more complex in construction as well as in operation because low power device has to change to make them appropriate in high power applications.

In power electronic circuits, this diode plays an essential role. It can be used as a rectifier in converter circuits, voltage regulation circuits, flyback / freewheeling diode, reverse voltage protection, etc.

These diodes are related to signal diodes except for a slight disparity in its construction. The doping level in signal diode for both P-layer & N-layer is the same whereas, in power diodes, the junction can be formed among a heavily doped P+ layer & lightly doped N– layer.

CONSTRUCTION

The construction of this diode includes three layers like the P+ layer, n– layer and n+ layer. Here the top layer is the P+ layer, it is heavily doped. The middle layer is n– layer, it is lightly doped and the last layer is n+ layer, and it is heavily doped.

Here p+ layer acts as an anode, the thickness of this layer is 10 µm. The n+ layer acts as a cathode, the thickness of this layer is 250-300 µm. The n- layer acts as a middle layer/drift layer, increases then breakdown voltage will be increased.

WORKING PRINCIPLE OF POWER DIODE

The working principle of this diode is similar to the normal PN junction diode. When the voltage of the anode terminal is high than the voltage of the cathode terminal, the diode conducts. The range of forwarding voltage drop in this diode is very small approximately 0.5V – 1.2V. In this mode, the diode works as a forward characteristic.

If the voltage of the cathode is high than the voltage of anode, the diode performs as blocking mode. In this mode, the diode performs like the reverse characteristic.

TYPES OF POWER DIODE

The power diodes depending on the reverse recovery time as well as the process of manufacturing are classified into three types such as

General Purpose Diodes
Fast Recovery Diodes
Schottky Diodes

APPLICATION OF POWER DIODE

Rectifiers
Clipper Circuits
Clamping Circuits
Reverse Current Protection Circuits
In Logic Gates

V-I CHARACTERISTICS OF POWER DIODES

The figure below shows the v-i characteristics of a power diode which is almost similar to that of a signal diode.

In signal diodes for forward biased region the current increases exponentially however in power diodes high forward current leads to high ohmic drop which dominates the exponential growth and the curve increases almost linearly. The maximum reverse voltage that the diode can withstand is depicted by V_{RRM}, i.e. peak reverse repetitive voltage. Above this voltage the reverse current becomes very high abruptly and as the diode is not designed to dissipate such high amount of heat, it may get destroyed. This voltage may also be called as peak inverse voltage (PIV).

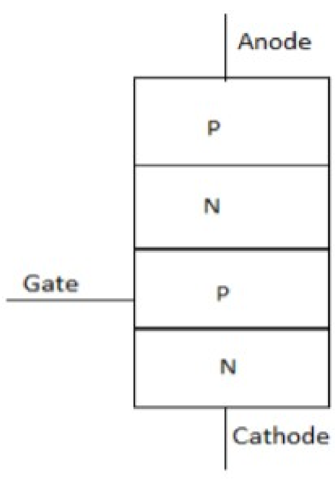
SCR/THYRISTOR

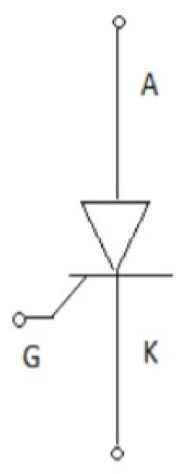
Thyristor is a four layer three junction pnpn semiconductor switching device. It has 3 terminals these are anode, cathode and gate. SCRs are solid state device, so they are compact, possess high reliability and have low loss.

CONSTRUCTION

An SCR is constructed with the four layers that consist of the P-type and the N-type semiconductor material. These are layered in such a way that it tends to form three junctions that are J1, J2, and J3. The three terminals that are attached to it are known as anode, cathode and the gate. The anode is the basic terminal through which the current flows or enters into the device. Where the cathode is the terminal through which the entered current leaves the device.

The current entering terminal is of positive polarity and the terminal through which the current is leaving is of negative polarity. In between the flow of current among the terminals, there must be a terminal that can provide the control. This can be provided by the terminal gate. This terminal is sometimes also referred to as the terminal of control.

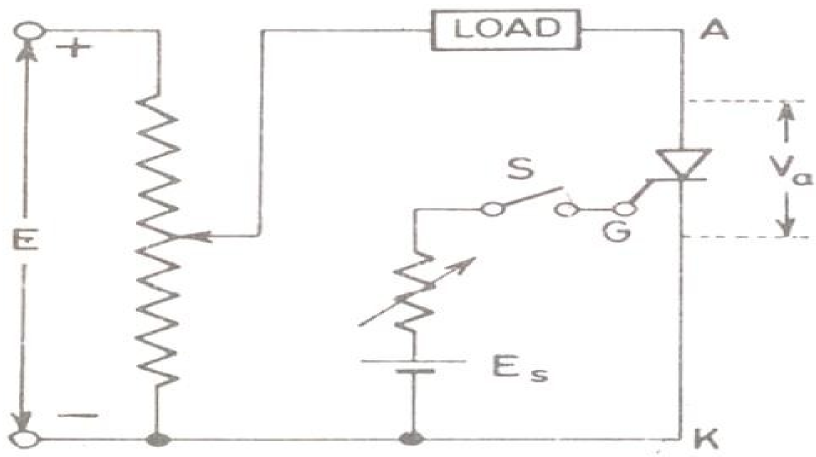




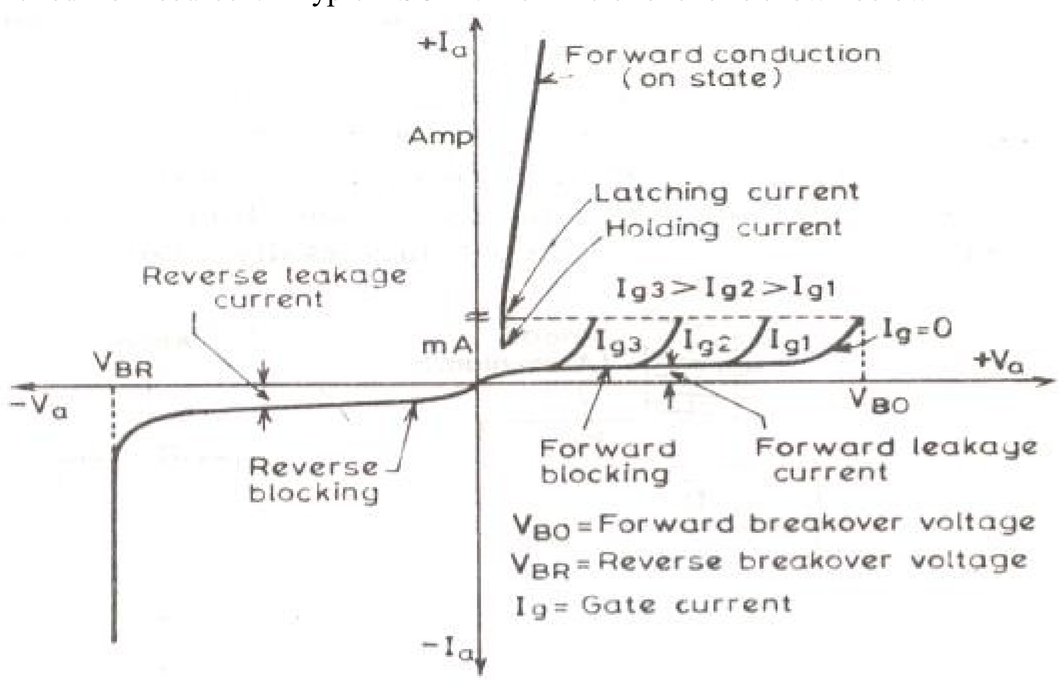
PRINCIPLE OF OPERATION

SCR is made up of silicon, it act as a rectifier; it has very low resistance in the forward direction and high resistance in the reverse direction. It is a unidirectional device.

The anode to cathode is connected in series with the load circuit. Essentially the device is a switch. Ideally it remains off (voltage blocking state), or appears to have an infinite impedance until both the anode and gate terminals have suitable positive voltages with respect to the cathode terminal. The thyristor then switches on and current flows and continues to conduct without further gate signals. Ideally the thyristor has zero impedance in conduction state. For switching off or reverting to the blocking state, there must be no gate signal and the anode current must be reduced to zero. Current can flow only in one direction.
The circuit diagram for obtaining static V-I characteristics is as shown



Anode and cathode are connected to main source voltage through the load. The gate and cathode are fed from source . A typical SCR V-I characteristic is as shown below



VBO=Forward breakover voltage
VBR=Reverse breakover voltage

Ig=Gate current
Va=Anode voltage across the thyristor terminal A,K.
Ia=Anode current
It can be inferred from the static V-I characteristic of SCR. SCR have 3 modes of operation:
1. Reverse blocking mode
2. Forward blocking mode (off state)
3. Forward conduction mode (on state)

1. Reverse Blocking Mode
When cathode of the thyristor is made positive with respect to anode with switch open thyristor is reverse biased. Junctions $J1$ and $J2$ are reverse biased where junction $J2$ is forward biased. The device behaves as if two diodes are connected in series with reverse voltage applied across them. A small leakage current of the order of few mA only flows. As the thyristor is reverse biased and in blocking mode. It is called as acting in reverse blocking mode of operation.
Now if the reverse voltage is increased, at a critical breakdown level called reverse breakdown voltage VBR, an avalanche occurs at $J1$ and $J3$ and the reverse current increases rapidly. As a large current associated with VBR and hence more losses to the SCR. This results in Thyristor damage as junction temperature may exceed its maximum temperature rise.

2. Forward Blocking Mode
When anode is positive with respect to cathode, with gate circuit open, thyristor is said to be forward biased.
Thus junction $J1$ and $J3$ are forward biased and $J2$ is reverse biased. As the forward voltage is increases junction $J2$ will have an avalanche breakdown at a voltage called forward breakover voltageVBO. When forward voltage is less then VBO thyristor offers high impedance. Thus a thyristor acts as an open switch in forward blocking mode.

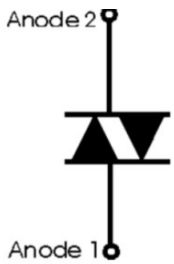
3. Forward Conduction Mode
Here thyristor conducts current from anode to cathode with a very small voltage drop across it. So a thyristor can be brought from forward blocking mode to forward conducting mode:
1. By exceeding the forward breakover voltage.
2. By applying a gate pulse between gate and cathode.
During forward conduction mode of operation thyristor is in on state and behave like a close switch. Voltage drop is of the order of 1 to 2V. This small voltage drop is due to ohmic drop across the four layers of the device.

LIST APPLICATIONS OF SCR
Due to the wide variety of advantages, like ability to turn ON from OFF state in response to a low gate current and also able to switch high voltages, makes the SCR or thyristor to be used in a variety of applications.

These applications include switching, rectification, regulation, protection, etc. The SCRs are used for home appliance control include lighting, temperature control, fan speed regulation, heating, and alarm activation.
For industrial applications, SCRs are used to control the motor speed, battery charging and power conversions.

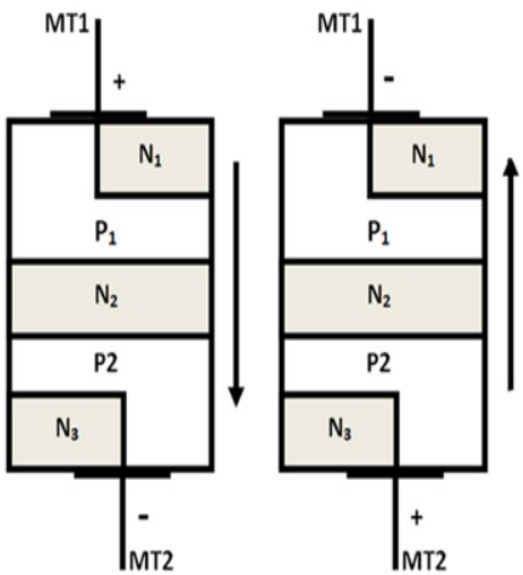
DIAC

The DIAC is a bi-directional semiconductor switch that can be switched on in both polarities. The full form of the name DIAC is diode alternating current.



CONSTRUCTION AND OPERATION OF DIAC

The basic construction of DIAC consist of two terminals namely MT1 and MT2. When the MT1 terminal is designed +Ve with respect to the terminal MT2, the transmission will take place to the p-n-p-n structure. The diac works in both the direction.



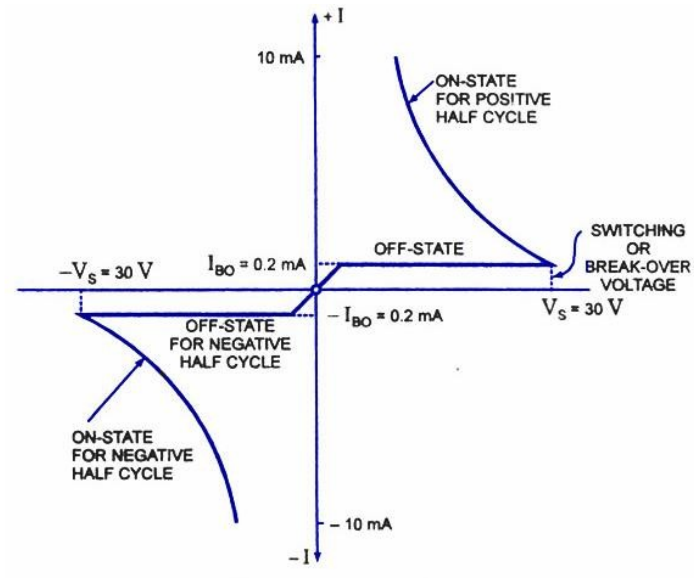
The DIAC is basically a diode that conducts after a 'break-over' voltage, selected VBO, and is exceeded. When the diode surpasses the break-over voltage, then it goes into the negative dynamic resistance of region. This causes in a reduce in the voltage drop across the diode with rising voltage. So there is a quick increase in the current level that is mannered by the device.
The diode leftovers in its transmission state until the current through it falls below, what is termed the holding current, which is usually chosen by the letters IH. Below the holding current,

the DIAC reverts to its non-conducting state. Its behavior is bidirectional and thus its function takes place on both halves of an alternating cycle.

CHARACTERISTICS OF DIAC

V-I characteristics of a diac is shown below

The diac performs like an open-circuit until its switching is exceeded. At that position the diac performs until its current decreases toward zero. Because of its abnormal construction, doesn't switch sharply into a low voltage condition at a low current level like the Triac or SCR, once it goes into transmission, the diac preserves an almost continuous –Ve resistance characteristic, that means, voltage reduces with the enlarge in current. This means that, unlike the Triac and the SCR, the diac cannot be estimated to maintain a low voltage drop until its current falls below the level of holding current.



V-I Characteristic of a Diac

LIST APPLICATION OF DIAC

The main application of a DIAC is its use in a TRIAC triggering circuit. The DIAC is connected to the gate terminal of the TRIAC. When the voltage across the gate decreases below a predetermined value, the gate voltage will be zero and hence the TRIAC will be turned off. Some other applications of a DIAC include:

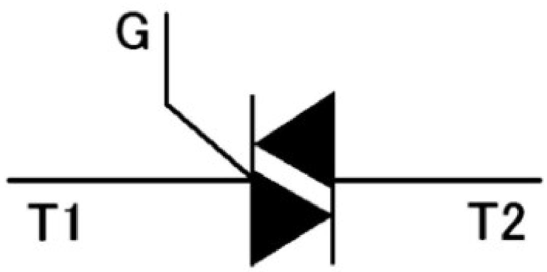
It can be used in the lamp dimmer circuit

It is used in a heat control circuit

It is used in the speed control of a universal motor

TRIAC

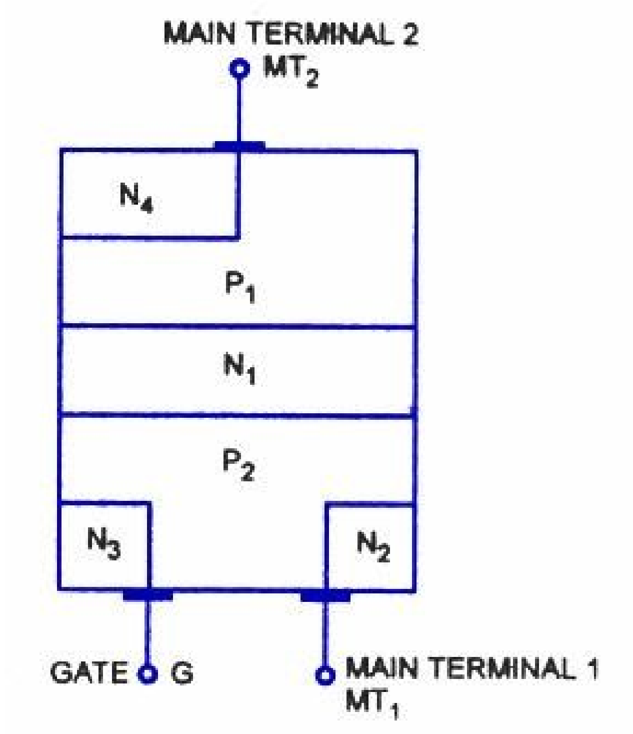
A Triac device comprises of two thyristors that are connected in opposite direction but in parallel but, it is controlled by the same gate. Triac is a 2-dimensional thyristor which is activated on both halves of the i/p AC cycle using + Ve or -Ve gate pulses. The three terminals of the Triac are MT1; MT2 & gate terminal (G). Generating pulses are applied between MT1 and gate terminals. The 'G' current to switch 100A from triac is not more than 50mA or so.



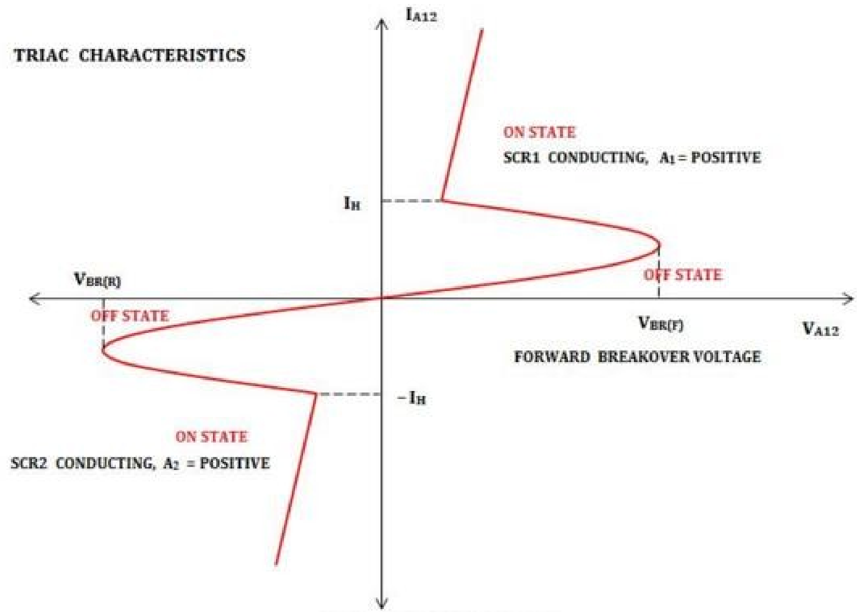
TRIAC

CONSTRUCTION AND OPERATION OF TRIAC

TRIAC is equivalent to two SCRs connected in inverse parallel with the gates connected together. As a result, the TRIAC functions as a Bidirectional switch to pass the current in both directions once the gate is triggered. TRIAC is a three terminal device with a Main terminal1 (MT1), Main terminal 2(MT2) and a Gate. The MT1 and MT2 terminals are used to connect the Phase and Neutral lines while the Gate is used to feed the triggering pulse. The Gate can be triggered either by a positive voltage or negative voltage. When the MT2 terminal gets a positive voltage with respect to the MT1 terminal and the Gate gets a positive trigger, then the left SCR of the TRIAC triggers and circuit completes. But if the polarity of the voltage at the MT2 and MT1 terminals is reversed and a negative pulse is applied to the Gate, then the right SCR of Triac conducts. When the Gate current is removed, the TRIAC switches off. So a minimum holding current Ih must be maintained at the gate to keep the TRIAC conducting.



Basic Structure



CHARACTERISTICS OF TRIAC

The V-I characteristics of TRIAC are discussed below

The triac is designed with two SCRs which are fabricated in the opposite direction in a crystal. Operating characteristics of triac in the 1st and 3rd quadrants are similar but for the direction of flow of current and applied voltage.

The V-I characteristics of triac in the first and third quadrants are basically equal to those of an SCR in the first quadrant.

It can be functioned with either +Ve or –Ve gate control voltage but in typical operation generally the gate voltage is +Ve in first quadrant and -Ve in third quadrant.

The supply voltage of the triac to switch ON depends upon the gate current. This allows utilizing a triac to regulate AC power in a load from zero to full power in a smooth and permanent manner with no loss in the device control.

MODES OF OPERATION OF TRIAC AND MENTION THE PREFERRED MODES

It is possible to connect various combinations of negative and positive voltages to the triac terminals because it is a bidirectional device. The four possible electrode potential combinations which make the triac to operate four different operating quadrants or modes are given as.

MT2 is positive with respect to MT1 with a gate polarity positive with respect to MT1.

MT2 is positive with respect to MT1 with a gate polarity negative with respect to MT1.

MT2 is negative with respect to MT1 with a gate polarity negative with respect to MT1.

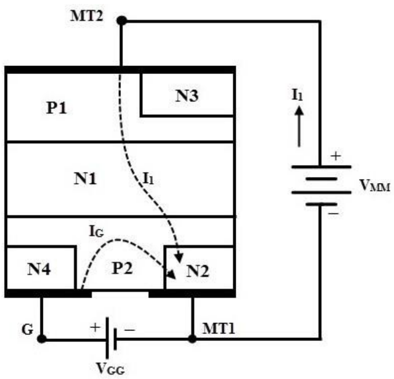
MT2 is negative with respect to MT1 with a gate polarity positive with respect to MT1.

Mode 1: MT2 is Positive, Positive Gate Current

When the gate terminal is made positive with respect to MT1, gate current flows through the P2 and N2 junction. When this current flows, the P2 layer is flooded with electrons and further these electrons are diffused to the edge of junction J2 (or P2-N1 junction).

These electrons collected by the N1 layer builds a space charge on the N1 layer. Therefore, more holes from the P1 region are diffused into the N1 region to neutralize the negative space charges. These holes arrive at the junction J2 and produce the positive space charge in the P2 region, which causes more electrons to inject into P2 from N2.

This results a positive regeneration and finally the main current flows from MT2 to MT1 through the regions P1- N1 – P2 – N2.

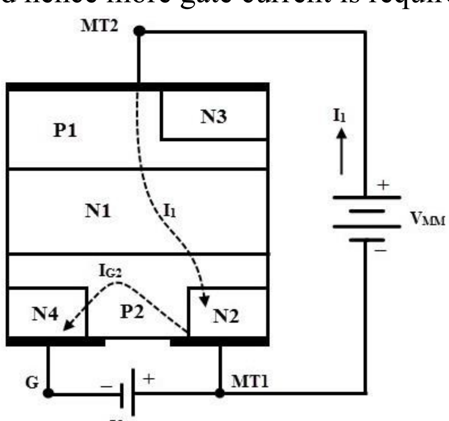


Mode 2: MT2 is Positive, Negative Gate Current

When MT2 is positive and the gate terminal is negative with respect to MT1, gate current flows through the P2-N4 junction. This gate current forward biases the P2-N4 junction for auxiliary P1N1P2N4 structure. This results the triac to conduct initially through the P1N1P2N4 layers.

This further raises the potential between P2N2 towards the potential of MT2. This causes the current to establish from left to right in the P2 layer which forward biases the junction P2N2. And hence the main structure P1N1P2N2 begins to conduct.

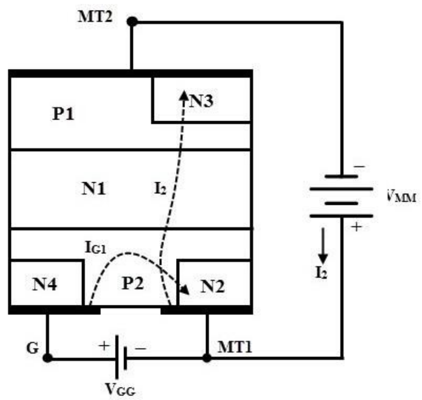
Initially conducted auxiliary structure P1N1P2N4 is considered as a pilot SCR while later conducted structure P1N1P2N2 is considered as main SCR. Hence the anode current of pilot SCR serves as gate current to the main SCR. The sensitivity to gate current is less in this mode and hence more gate current is required to turn the triac.



Mode 3: MT2 is Negative, Positive Gate Current

In this mode, MT2 is made negative with respect to MT1 and the device is turned ON by applying a positive voltage between the gate and MT1 terminal. The turn ON is initiated by N2 which acts as a remote gate control and the structure leads to turn ON the triac is P2N1P1N3.

The external gate current forward biases the junction P2-N2. N2 layer injects the electrons into the P2 layer which are then collected by junction P2N1. This result to increases the current flow through P2N1 junction.



The holes injected from layer P2 diffuse through the N1 region. This builds a positive space charge in the P region. Therefore, more electrons from N3 are diffused into P1 to neutralize the positive space charges.

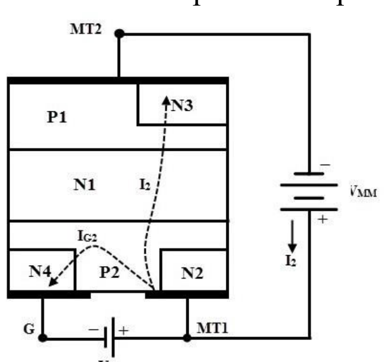
Hence, these electrons arrive at junction J2 and produce a negative space charge in the N1region which results to inject more holes from the P2 into the region N1. This regenerative process continues till the structure P2N1P1N3 turns ON the triac and conducts the external current.

As the triac is turned ON by the remote gate N2, the device is less sensitive to the positive gate current in this mode.

Mode 4: MT2 is Negative, Negative Gate Current

In this mode N4 acts as a remote gate and injects the electrons into the P2 region. The external gate current forward biases the junction P2N4. The electrons from the N4 region are collected by the P2N1 junction increase the current across P1N1 junction.

Hence the structure P2N1P1N3 turns ON by the regenerative action. The triac is more sensitive in this mode compared with positive gate current in mode 3.



From the above discussion, it is concluded that the modes 2 and 3 are less sensitive configuration which needs more gate current to trigger the triac, whereas more common triggering modes of triac are 1 and 4 which have greater sensitivity. In practice the more sensitive mode of operation is selected such that the polarity of the gate is to match with the polarity of the terminal MT2.

APPLICATIONS OF TRIAC:

TRIACs are used in numerous applications such as light dimmers, speed controls for electric fans and other electric motors and in the modern computerized control circuits of numerous household small and major appliances. They can be used both into AC and DC circuits however the original design was to replace the utilization of two SCRs in AC circuits. There are two families of TRIACs, which are mainly used for application purpose, they are BT136, BT139.

POWER MOSFET

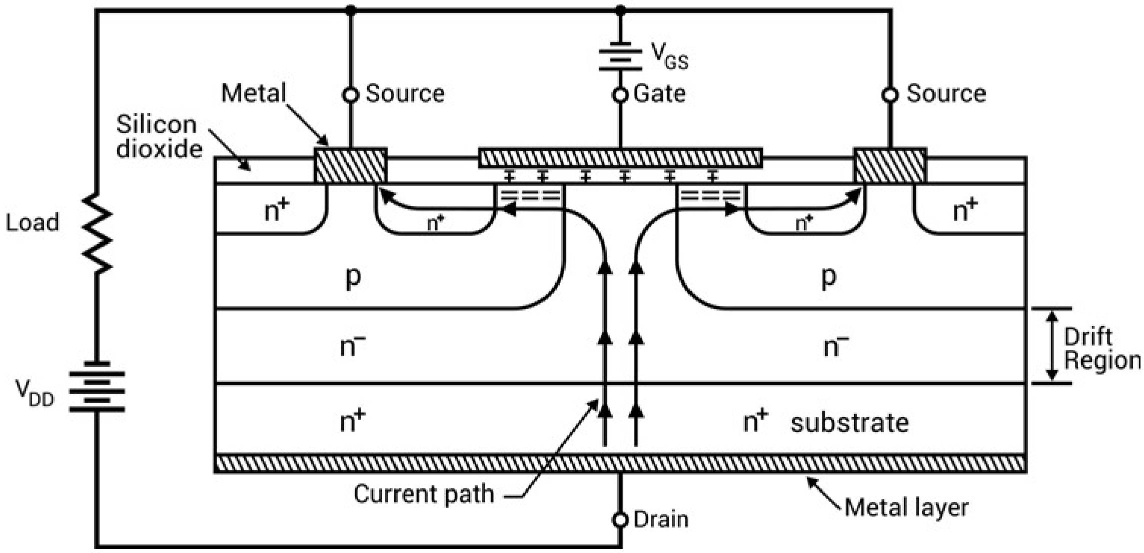
The MOSFET (Metal Oxide Semiconductor Field Effect Transistor) transistor is a semiconductor device that is widely used for switching purposes and for the amplification of electronic signals in electronic devices.

The constructional details of high power MOSFET are shown in below figure. In this figure is shown a planar diffused metal-oxide-semiconductor (DMOS) structure for n -channel which is quite common for power MOSFETs. On n+ substrate, high resistivity n− layer is epitaxially grown. The thickness of n− layer determines the voltage blocking capability of the device.

On the other side of n+ substrate, a metal layer is deposited to form the drain terminal. Now p− regions are diffused in the epitaxially grown n− layer. Further, n+ regions are diffused in p regions as shown. As before, Si02 layer is added, which is then etched so as to fit metallic source and gate terminals.

A power MOSFET actually consists of a parallel connection of thousands of basic MOSFET cells on the same single chip of silicon.
When gate circuit voltage is zero, and VDD is present, n−p junctions are reverse biased and no current flows from drain to source. When gate terminal is made positive with respect to source, an electric field is established and electrons form n-channel in the p regions as shown. So a current from drain to source is established as indicated by arrows.

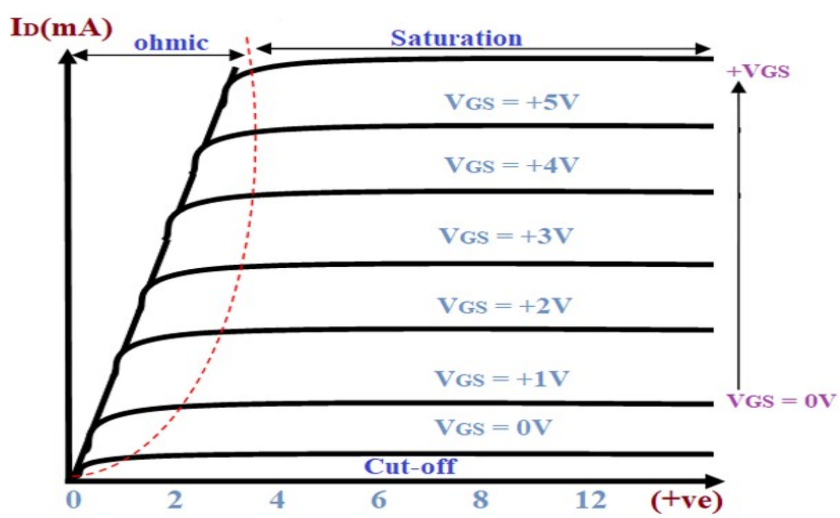
With gate voltage increased current ID also increases as expected. Length of n-channel can be controlled and therefore on-resistance can be made low if short length is used for the channel.
Power MOSFET conduction is due to majority carriers; therefore, time delays caused by removal or recombination of minority carriers are eliminated. Thus, power MOSFET can work at switching frequencies in the megahertz range.



Basic structure of a n-channel DMOS power MOSFET

VI Characteristics:

VI characteristics of the enhancement-mode MOSFET are drawn between the drain current (I_D) and the drain-source voltage (V_{DS}). The VI characteristics are partitioned into three different regions, namely ohmic, saturation, and cut-off regions. The cutoff region is the region where the MOSFET will be in the OFF state where the applied bias voltage is zero. When the bias voltage is applied, the MOSFET slowly moves towards conduction mode, and the slow increase in conductivity takes place in the ohmic region. Finally, the saturation region is where the positive voltage is applied constantly and the MOSFET will be staying in the conduction state.

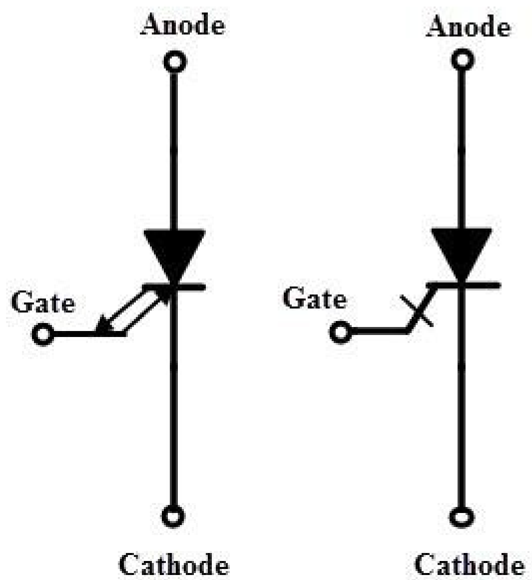


APPLICATIONS OF MOSFET

- MOSFET amplifiers are extensively used in radio frequency applications.
- It acts as a passive element like resistor, capacitor and inductor.
- DC motors can be regulated by power MOSFETs.
- High switching speed of MOSFETs makes it an ideal choice in designing chopper circuits.

GTO

A Gate Turn off Thyristor or GTO is a three terminal, bipolar (current controlled minority carrier) semiconductor switching device. Similar to conventional thyristor, the terminals are anode, cathode and gate as shown in figure below. As the name indicates, it has gate turn off capability. These are capable not only to turn ON the main current with a gate drive circuit, but also to turn it OFF.

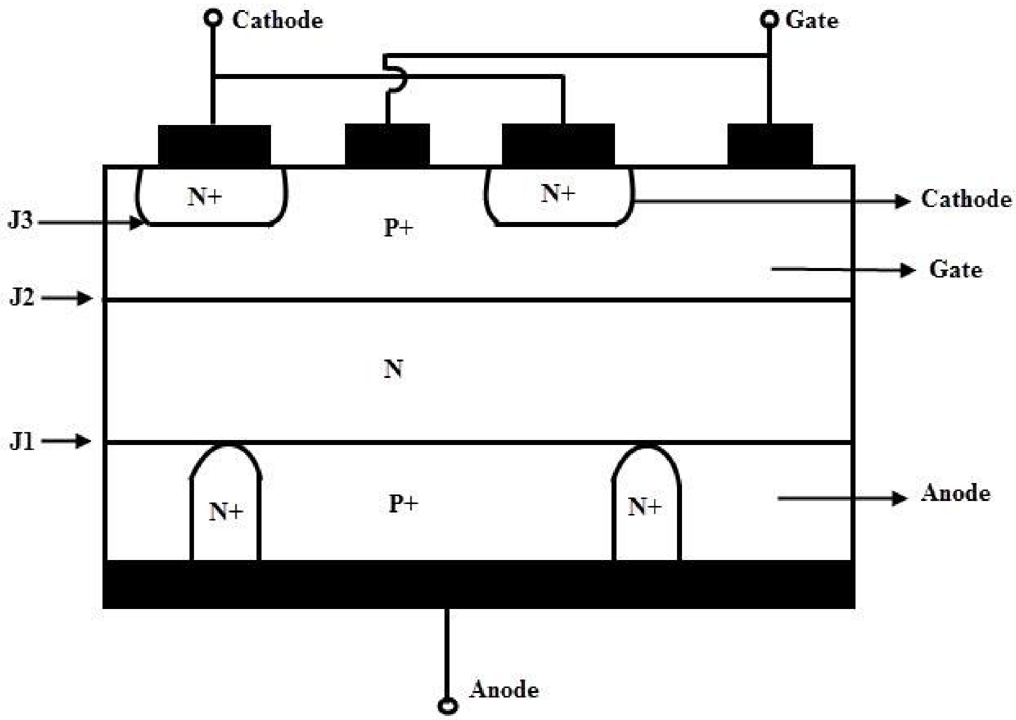


Gate Turn-Off Thyristor Symbols

CONSTRUCTION

Consider the below structure of GTO, which is almost similar to the thyristor. It is also a four layer, three junction P-N-P-N device like a standard thyristor. In this, the n+ layer at the cathode end is highly doped to obtain high emitter efficiency. This result the breakdown voltage of the junction J3 is low which is typically in the range of 20 to 40 volts.

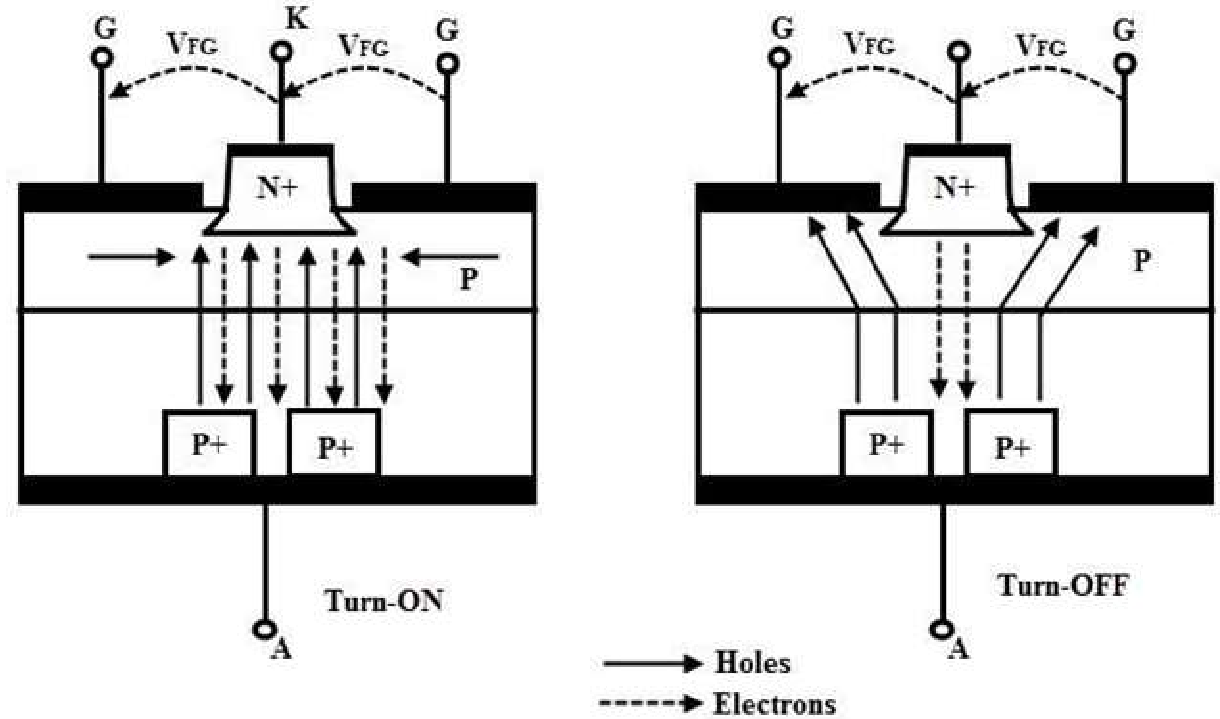
The doping level of the p type gate is highly graded because the doping level should be low to maintain high emitter efficiency, whereas for having a good turn OFF properties, doping of this region should be high. In addition, gate and cathodes should be highly interdigited with various geometric forms to optimize the current turn off capability.



The junction between the P+ anode and N base is called anode junction. A heavily doped P+ anode region is required to obtain the higher efficiency anode junction so that a good turn ON properties is achieved. However, the turn OFF capabilities are affected with such GTOs.
This problem can be solved by introducing heavily doped N+ layers at regular intervals in P+ anode layer as shown in figure. So this N+ layer makes a direct contact with N layer at junction J1. This cause the electrons to travel from base N region directly to anode metal contact without causing hole injection from P+ anode. This is called as a anode shorted GTO structure.
Due to these anode shorts, the reverse blocking capacity of the GTO is reduced to the reverse breakdown voltage of junction j3 and hence speeds up the turn OFF mechanism.
However, with a large number of anode shorts, the efficiency of the anode junction reduces and hence the turn ON performance of the GTO degrades. Therefore, careful considerations have to be taken about the density of these anode shorts for a good turn ON and OFF performance.

PRINCIPLE OF OPERATION
The turn ON operation of GTO is similar to a conventional thyristor. When the anode terminal is made positive with respect to cathode by applying a positive gate current, the hole current injection from gate forward bias the cathode p-base junction.
This results in the emission of electrons from the cathode towards the anode terminal. This induces the hole injection from the anode terminal into the base region. This injection of holes and electrons continuous till the GTO comes into the conduction state.
In case of thyristor, the conduction starts initially by turning ON the area of cathode adjacent to the gate terminal. And thus, by plasma spreading the remaining area comes into the conduction.

Unlike a thyristor, GTO consists of narrow cathode elements which are heavily interdigitated with gate terminal, thereby initial turned ON area is very large and plasma spreading is small. Hence the GTO comes into the conduction state very quickly.



To turn OFF a conducting GTO, a reverse bias is applied at the gate by making the gate negative with respect to cathode. A part of the holes from the P base layer is extracted through the gate which suppress the injection of electrons from the cathode.

In response to this, more hole current is extracted through the gate results more suppression of electrons from the cathode. Eventually, the voltage drop across the p base junction causes to reverse bias the gate cathode junction and hence the GTO is turned OFF.

During the hole extraction process, the p-base region is gradually depleted so that the conduction area squeezed. As this process continuous, the anode current flows through remote areas forming high current density filaments. This causes local hot spots which can damage the device unless these filaments are extinguished quickly.

By the application of high negative gate voltage these filaments are extinguished rapidly. Due to the N base region stored charge, the anode to gate current continues to flow even though the cathode current is ceased. This is called a tail current which decays exponentially as the excess charge carriers are reduced by the recombination process. Once the tail current reduced to a leakage current level, the device retains its forward blocking characteristics.

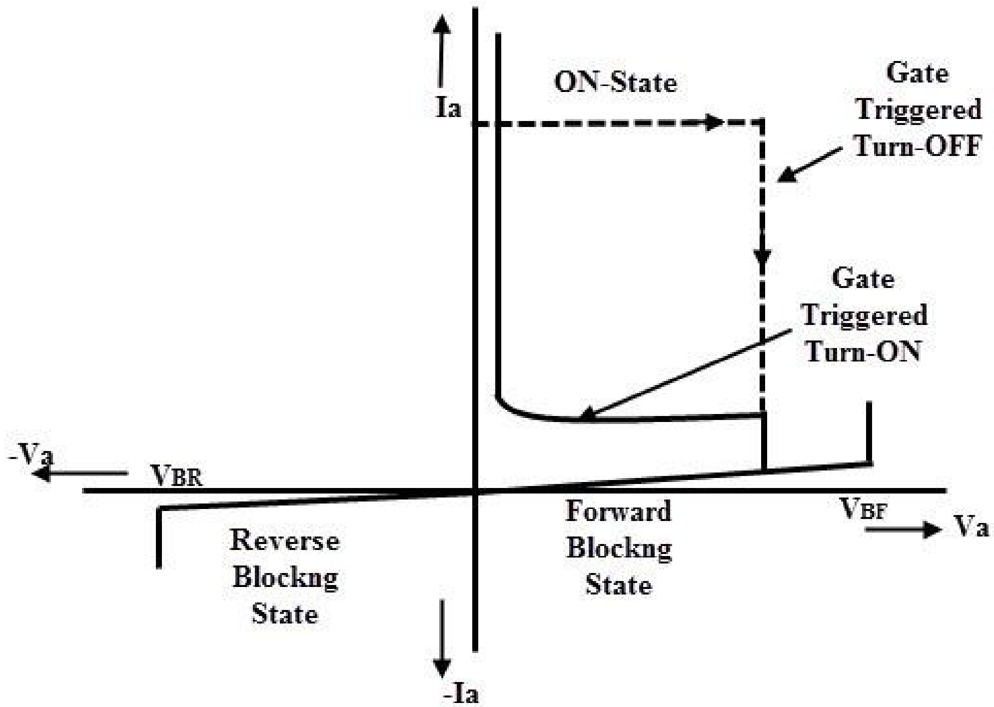
V-I CHARACTERISTICS

During the turn ON, GTO is similar to thyristor in its operates.So the first quadrant characteristics are similar to the thyristor. When the anode is made positive with respect to cathode, the device operates in forward blocking mode. By the application of positive gate signal triggers the GTO into conduction state.

The latching current and forward leakage currents are considerably higher in GTO compared to the thyristor as shown in figure. The gate drive can be removed if the anode current is above the holding current level.

But it is recommended not to remove the positive gate drive during conduction and to hold at value more than the maximum critical gate current. This is because the cathode is subdivided into small finger elements as discussed above to assist the turn OFF process.

This causes the anode current dips below the holding current level transiently, which forces a high anode current at a high rate back into the GTO. This can be potentially destructive. Therefore, some manufacturers recommend the continuous gate signal during the conduction state.



The GTO can be turned OFF by the application of reverse gate current which can be either step or ramp drive. The GTO can be turned OFF without reversing anode voltage. The dashed line in the figure shows i-v trajectory during the turn OFF for an inductive load. It should be noted that during the turn OFF, GTO can block a rated forward voltage only.

To avoid dv/dt triggering and protect the device during turn OFF, either a recommended value of resistance must be connected between the gate and cathode or a small reverse bias voltage (typically -2V) must be maintained on the gate terminal. This prevents the gate cathode junction to become forward biased and hence the GTO sustains during the turn OFF state.

In reverse biased condition of GTO, the blocking capability is depends on the type of GTO. A symmetric GTO has a high reverse blocking capability while asymmetric GTO has a small reverse blocking capability as shown in figure.

It is observed that, during reverse biased condition, after a small reverse voltage (20 to 30 V) GTO starts conducting in reverse direction due to the anode short structure. This mode of operation does not destroy the device provided that the gate is negatively biased and the time of this operation should be small.

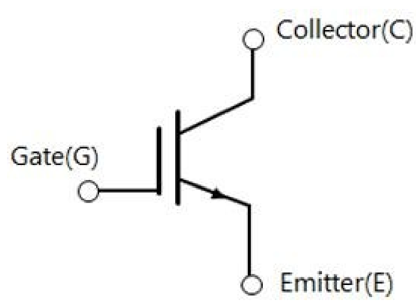
LIST APPLICATION OF GTO

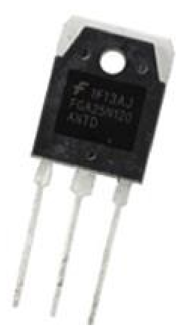
Due to the advantages like excellent switching characteristics, no need of commutation circuit, maintenance-free operation, etc makes the GTO usage predominant over thyristor in many applications. It is used as a main control device in choppers and inverters. Some of these applications are

- AC drives
- DC drives or DC choppers
- AC stabilizing power supplies
- DC circuit breakers
- Induction heating
- And other low power applications

IGBT

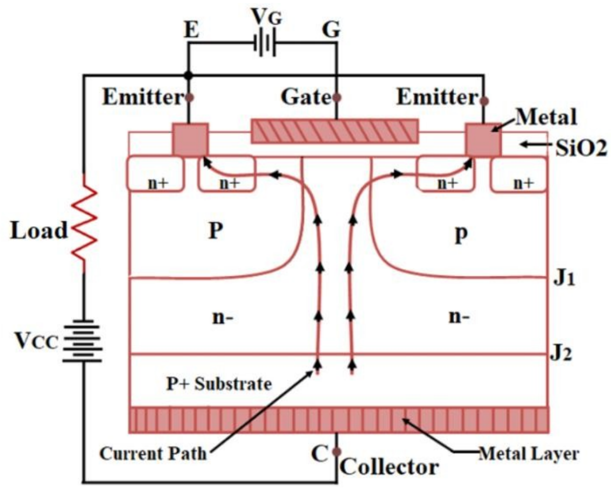
IGBT is the short form of Insulated Gate Bipolar Transistor. It is a three-terminal semiconductor switching device that can be used for fast switching with high efficiency in many types of electronic devices. These devices are mostly used in amplifiers for switching/processing complex wave patters with pulse width modulation (PWM). The typical symbol of IGBT along with its image is shown below.





WORKING OF IGBT

IGBT has three terminals attached to three different metal layers, the metal layer of the gate terminal is insulated from the semiconductors by a layer of silicon dioxide (SIO2). IGBT is constructed with 4 layers of semiconductor sandwiched together. The layer closer to the collector is the p+ substrate layer above that is the n- layer, another p layer is kept closer to the emitter and inside the p layer, we have the n+ layers. The junction between the p+ layer and n- layer is called the junction J2 and the junction between the n- layer and the p layer is called the junction J1. The structure of IGBT is shown in the figure below.

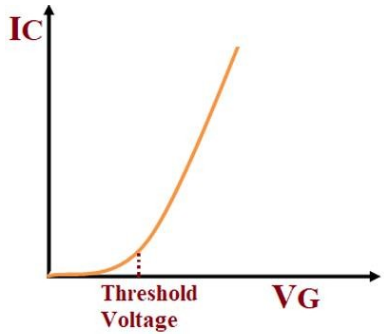


To understand the working of the IGBT, consider a voltage source VG connected positively to the Gate terminal with respect to the Emitter. Consider other voltage source VCC connected across The Emitter and the Collector, where Collector is kept positive with respect to the Emitter. Due to the voltage source VCC the junction J1 will be forward-biased whereas the junction J2 will be reverse biased. Since J2 is in reverse bias there will not be any current flow inside the IGBT (from collector to emitter).

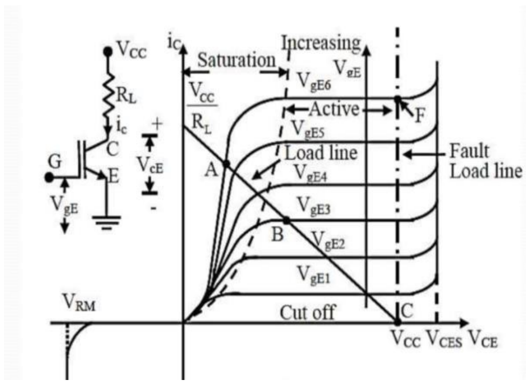
Initially, consider that there is no voltage applied to the Gate terminal, at this stage the IGBT will be in a non-conductive state. Now if we increase the applied gate voltage, due to the capacitance effect on the SiO2 layer the negative ions will get accumulated on the upper side of the layer and the positive ions will get accumulated on the lower side of the SiO2 layer. This will cause the insertion of negative charge carriers in the p region, higher the applied voltage VG greater the insertion of negatively charged carriers. This will lead to a formation of the channel between the J2 junctions which allow the flow of current from collector to emitter. The flow of current is represented as the current path in the picture, when the applied Gate voltage VG increases the amount of current flow from the collector to the emitter also increases.

INPUT CHARACTERICTCS OF IGBT

The input characteristics of IGBT can be understood from the graph below. Initially, when no voltage is applied to the gate pin the IGBT is in turn off condition and no current flows through the collector pin. When the voltage applied to the gate pin exceeds the threshold voltage, the IGBT starts conducting and the collector current IG starts to flow between the collector and emitter terminals. The collector current increases with respect to the gate voltage as shown in the graph below.



VI CHARACTRSTICS OF IGBT



The output characteristics of IGBT

The output characteristics of IGBT have three stages, initially, when the Gate Voltage V_{GE} **is** zero the device is in the off state, this is called the cutoff region**.** When V_{GE} is increased and if it is less than the threshold voltage then there will be a small leakage current flowing through the device, but the device will still be in the cutoff region. When the V_{GE} is increased beyond the threshold voltage the device goes into the active region and the current starts flowing through the device. The flow of current will increases with an increase in the voltage V_{GE} as shown in the graph above. The output characteristics of IGBT have three stages, initially, when the Gate Voltage VGE is zero the device is in the off state, this is called the cutoff region. When VGE is increased and if it is less than the threshold voltage then there will be a small leakage current flowing through the device, but the device will still be in the cutoff region. When the VGE is increased beyond the threshold voltage the device goes into the active region and the current starts flowing through the device. The flow of current will increases with an increase in the voltage VGE as shown in the graph above.

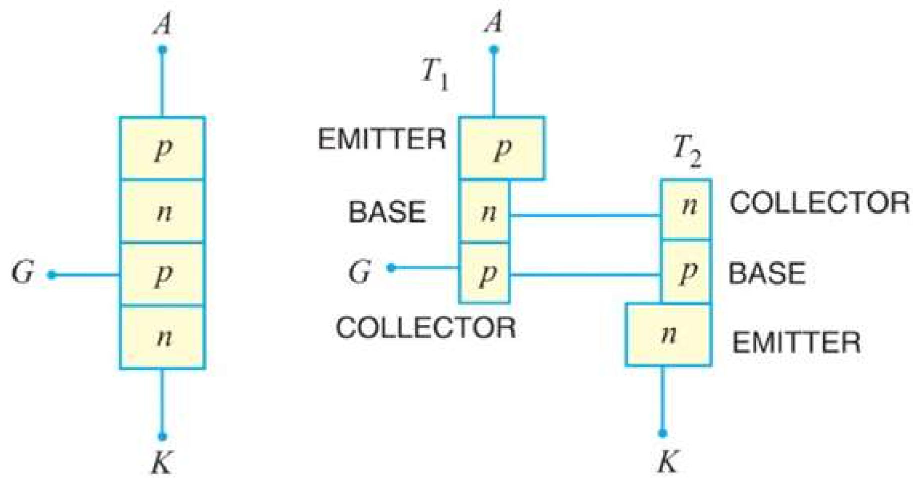
APPLICATIONS OF IGBT

IGBTs are used in various applications such as AC and DC motor drives, Unregulated Power Supply (UPS), Switch Mode Power Supplies (SMPS), traction motor control and induction heating, inverters, used to combine an isolated-gate FET for the control input and a bipolar power transistor as a switch in a single device, etc.

TWO TRANSISTOR ANALOGY OF SCR.

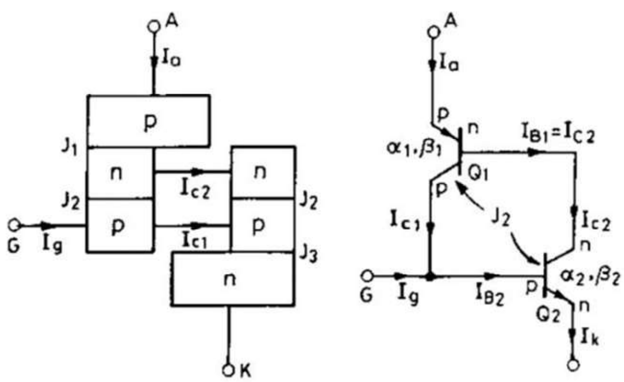
Two transistor analogy of SCR is a method of representing SCR in the form of two transistor model. This represents SCR is the combination of PNP and NPN transistor.

SCR or thyristor is a three terminal semiconductor device which having P-N-P-N structure. The basic operating principle of SCR can understand by two transistor method of SCR.



As per figure you can see **two transistors equivalent circuit of SCR**. From the figure, you can see the base of the transistor T1 is work as the collector of the transistor T2 and collector of the transistor T1 work as the base of the transistor T2.

Now here we find the expression for anode current of SCR.



As per transistor leakage current equation,

Collector current is expressed as,

$$I_C = \alpha I_E + I_{CBO}$$

Where α is the current gain of transistor and Icbo is the leakage current of the common base transistor.

For transistor T1 emitter current = anode current Ia and collector current Ic = Ic1

$$I_{C1} = \alpha_1 I_a + I_{CBO1}$$

Where α1 is the current gain of transistor T1.

Similarly, for transistor T2

$$I_{C2} = \alpha_2 I_k + I_{CBO2}$$

Where α2 is the current gain of transistor T2. And emitter current of transistor T2 = cathode current Ik.

Hereby figure, you can see anode current Ia is the sum of two collector current: Ic1 and Ic2.

$$\therefore \quad I_a = I_{C1} + I_{C2}$$
$$I_a = \alpha_1 I_a + I_{CBO1} + \alpha_2 I_k + I_{CBO2}$$

By putting Ik = Ia + Ig, anode current Ia will be,

$$I_a = \alpha_1 I_a + I_{CBO1} + \alpha_2 (I_a + I_g) + I_{CBO2}$$
$$I_a = \frac{\alpha_2 I_g + I_{CBO1} + I_{CBO2}}{1 - (\alpha_1 + \alpha_2)}$$

From this relation we can assure that with increasing the value of $(\alpha_1 + \alpha_2)$ towards unity, corresponding anode current will increase.

GATE CHARACTERISTICS OF THYRISTOR

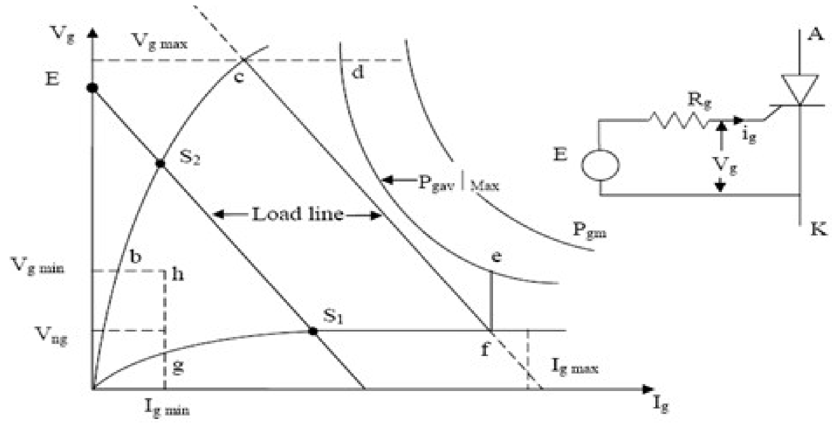
The gate circuit of a thyristor behaves like a poor quality diode with high on state voltage drop and low reverse break down voltage. This characteristic usually is not unique even within the same family of devices and shows considerable variation from device to device.

Therefore, manufacturer's data sheet provides the upper and lower limit of this characteristic as shown in figure below. Each thyristor has maximum gate voltage limit (Vgmax), gate current limit (Igmax) and maximum average gate power dissipation limit (PGav/Max). These limits should not be exceeded in order to avoid permanent damage to the gate cathode junction.

There are also minimum limits of Vg (Vgmin) and Ig (Igmin) for reliable turn on of the thyristor. A gate non triggering voltage (Vng) is also specified by the manufacturers of Thyristors.

All spurious noise signals should be less than this voltage Vng in order to prevent unwanted turn on of the thyristor.

Referring to the gate drive circuit in the inset the equation of the load line is given by Vg = E – Rgig



Gate Drive Requirements

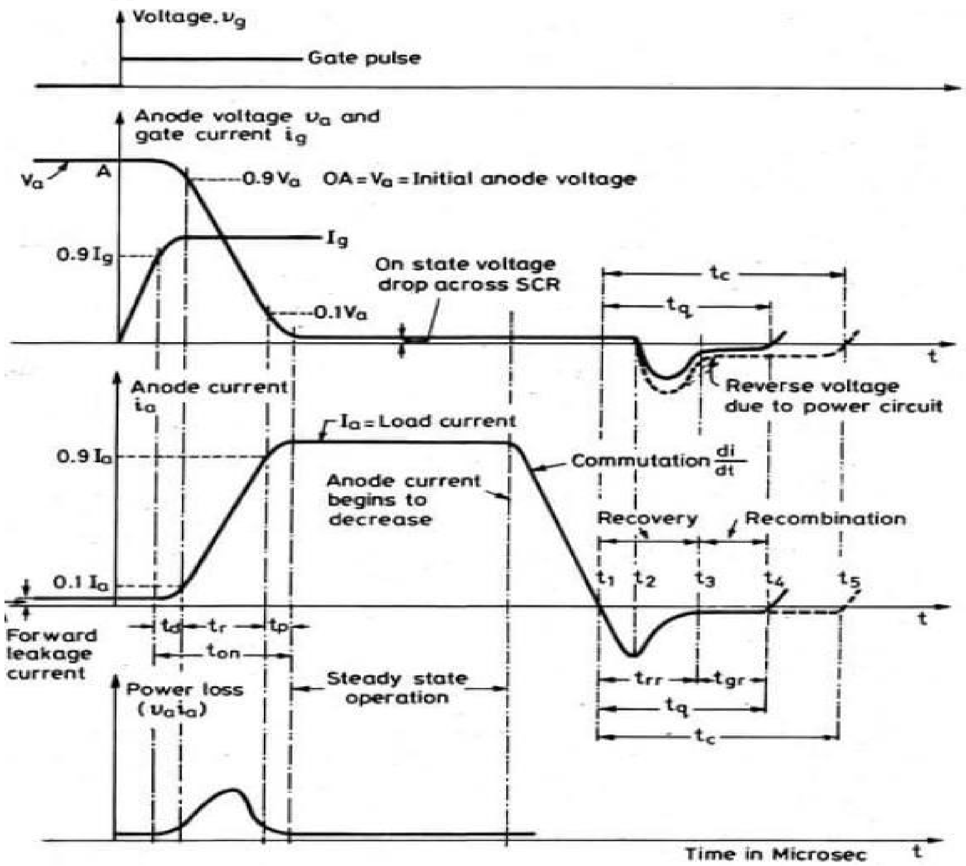
- Positive gate voltage or gate current
- Maximum Permissible gate power dissipation P_{GM}
- There are maximum and minimum limits for gate voltage and gate current to prevent the permanent destruction of junction J3 and to provide the realizable triggering.
- The gate signal can be ac or dc or a sequence of high frequency pulses.

SWITCHING CHARACTERISTIC OF SCR DURING TURN ON AND TURN OFF. The device's switching characteristics tells us about the switching losses, which is veryimportant parameter to decide the selection of device.

- When a positive gate signal is applied to a forward biased SCR, the transition of SCR from blocking state to conducting state is called as turn ON mechanism.
- The time taken for SCR to traverse from the blocking state to conducting state is called as turn on time.
- Turn on time is divided into 3 periods.
- tON = td + tr + tp
- td = delay time, tp or ts = peak time (or) spread time
- when the gate current reaches 0.9IG the anode current IA starts increasing and reaches 0.1IA (10% of its max value)
- The time taken for anode current to reach 0.1IA is called as delay time (td).
- In other words, it is the time taken for anode voltage to fall from VA to 0.9VA
- The anode current further increases and reaches 0.9IA.
- The time taken by the anode current to increases from 0.1IA to 0.9IA is called as rise time (tr).
- In other words, it is the time taken by the anode voltage to fall from 0.9VA to 0.1VA
- Spread Time or Peak time (ts or tp)
- It is time taken by the anode current to rise from (0.9IA to maximum value of IA) 90% to 100% of its full value. (or)
- It is the time taken by VA to fall from 0.1VA to its ON state voltage drop (near by zero).
- During this time the conduction spreads over the entire cross-section of cathode and so electrons spread over all the junctions.

TURN OFF MECHANISM:

- Turning OFF an SCR means bringing the SCR from conducting state to blocking state.
- To turn off an SCR two things are to be done
- Reduce the anode current below its holding current level.
- Application of reverse voltage.
- When the anode current is zero, if we apply forward voltage to the SCR, the device will not be able to block this forward voltage due to the fact that excess charge carriers are still at the junctions, so the device will start conducting even when the gate signal is not applied



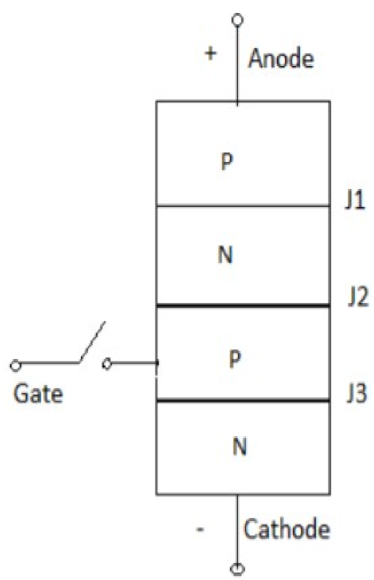
- In order to avoid this, reverse biasing of SCR is done to remove the excess charge carriers from all four layers.
- The turn OFF time is defined as the time from the instant the anode current becomes zero to the instant SCR reaches its forward blocking ability.
- Turn off time tOFF = trr + tgr
- trr = Reverse recovery time
- tgr =Gate recovery time
- Reverse recovery process is the removal of excessive charge carries from the top and bottom layers of SCR.
- At t1; current IA = 0

- After t1; IA build up in the reverse direction, due to the charge carriers stored in the four layers.
- Reverse recovery current removes the excessive carriers from junctions J1 and J3 during the time t1 to t3. (Reverse recovery current flows due sweeping out of holes from top p-layer and electrons from bottom n layer)
- Reverse Recovery Time (trr):-
- It is the time taken for the removal of excessive carriers from top and bottom layer of SCR.
- At t2: When nearly 60% of charges are removed from the outer two layers, the reverse recovery current decreases.
- This decaying causes a reverse voltage to be applied across the SCR.
- At t3 all excessive carriers from J1 and J3 is removed.
- The reverse voltage across SCR removes the excessive carriers from junction J2.
- Gate recovery process is the removal of excessive carriers from J2 junction by application of reverse voltage.
- Time taken for removal of trapped charges from J2 is called gate recovery time(tgr).
- At t4 all the carriers are removed and the device moves to the forward blocking mode.

DIFFERENT TURN ON METHODS FOR SCR

1. Forward voltage triggering
2. Gate triggering
3. $dv\ dt$ triggering
4. Light triggering
5. Temperature triggering

1. FORWARD VOLTAGE TRIGGERING



A forward voltage is applied between anode and cathode with gate circuit open.

- Junction $J1$ and $J3$ is forward biased
- Juntion $J2$ is reverse biased.

- As the anode to cathode voltage is increased breakdown of the reverse biased junction $J2$ occurs. This is known as avalanche breakdown and the voltage at which this phenomena occurs is called forward breakover voltage.
- The conduction of current continues even if the anode cathode voltage reduces below VBO till Ia will not go belowIh. Where Ih is the holding current for the thyristor.

2. GATE TRIGGERING

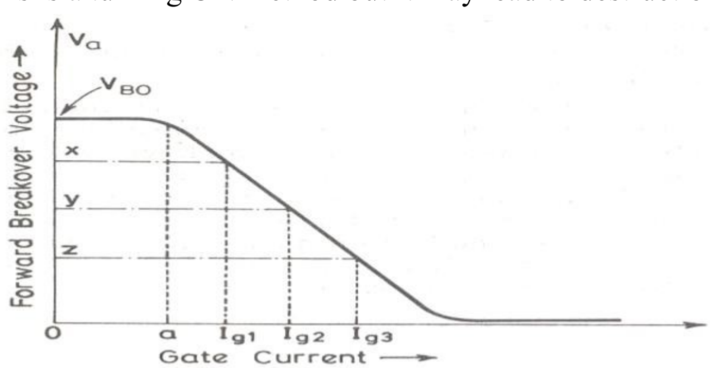
This is the simplest, reliable and efficient method of firing the forward biased SCRs. First SCR is forward biased. Then a positive gate voltage is applied between gate and cathode. In practice the transition from OFF state to ON state by exceeding VBO is never employed as it may destroy the device. The magnitude of VBO, so forward breakover voltage is taken as final voltage rating of the device during the design of SCR application.

First step is to choose a thyristor with forward breakover voltage (say 800V) higher than the normal working voltage. The benefit is that the thyristor will be in blocking state with normal working voltage applied across the anode and cathode with gate open. When we require the turning ON of a SCR a positive gate voltage between gate and cathode is applied. The point to be noted that cathode n- layer is heavily doped as compared to gate p-layer. So when gate supply is given between gate and cathode gate p-layer is flooded with electron from cathode n-layer. Now the thyristor is forward biased, so some of these electron reach junction $J2$ .As a result width of $J2$ breaks down or conduction at $J2$ occur at a voltage less than VBO.As Ig increases VBO reduces which decreases then turn ON time. Another important point is duration for which the gate current is applied should be more then turn ON time. This means that if the gate current is reduced to zero before the anode current reaches a minimum value known as holding current, SCR can't turn ON.

In this process power loss is less and also low applied voltage is required for triggering.

3. DV/DT TRIGGERING

This is a turning ON method but it may lead to destruction of SCR and so it must be avoided



When SCR is forward biased, junction $J1$ and $J3$ are forward biased and junction $J2$ is reversed biased so it behaves as if an insulator is place between two conducting plate. Here $J1$ and $J3$ acts as a conducting plate and $J2$ acts as an insulator. $J2$ is known as junction capacitor. So if we increase the rate of change of forward voltage instead of increasing the magnitude of voltage.

Junction $J2$ breaks and starts conducting. A high value of changing current may damage the SCR. So SCR may be protected from high dv/ dt.

$q = cv$

$Ia = c\ dv/dt$

$Ia\ \alpha\ dv/dt$

4. TEMPERATURE TRIGGERING

During forward biased, $J2$ is reverse biased so a leakage forward current always associated with SCR. Now as we know the leakage current is temperature dependant, so if we increase the temperature the leakage current will also increase and heat dissipitation of junction $J2$occurs. When this heat reaches a sufficient value $J2$ will break and conduction starts.

Disadvantages

This type of triggering causes local hot spot and may cause thermal run away of the device. This triggering cannot be controlled easily. It is very costly as protection is costly.

5. Light triggering

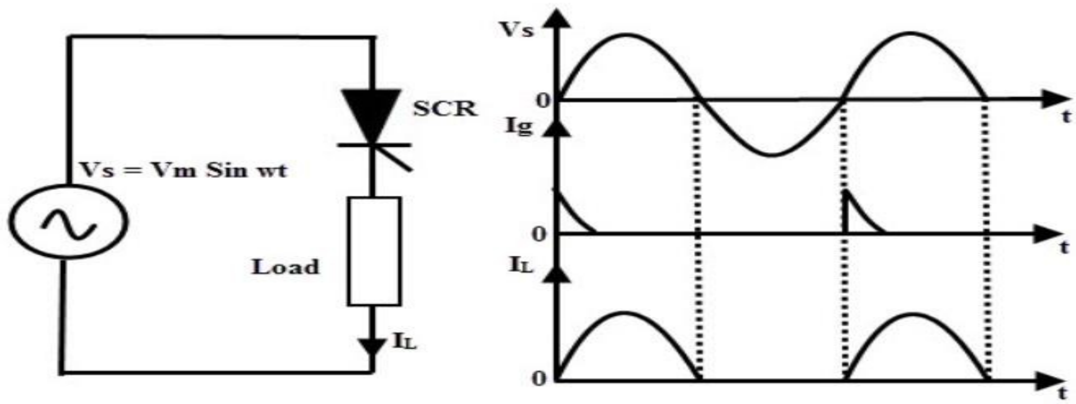
First a new recess niche is made in the inner p-layer. When this recess is irradiated, then free charge carriers (electron and hole) are generated. Now if the intensity is increased above a certain value then it leads to turn ON of SCR. Such SCR are known as Light activated SCR (LASCR). Some definitions: Latching current The latching current may be defined as the minimum value of anode current which at must attain during turn ON process to maintain conduction even if gate signal is removed. Holding current It is the minimum value of anode current below which if it falls, the SCR will turn OFF.

TURN OFF METHODS OF SCR (LINE COMMUTATION AND FORCED COMMUTATION)

NATURAL COMMUTATION/ **LINE COMMUTATION**

In natural commutation, the source of commutation voltage is the supply source itself. If the SCR is connected to an AC supply, at every end of the positive half cycle the anode current goes through the natural current zero and also immediately a reverse voltage is applied across the SCR. These are the conditions to turn OFF the SCR.

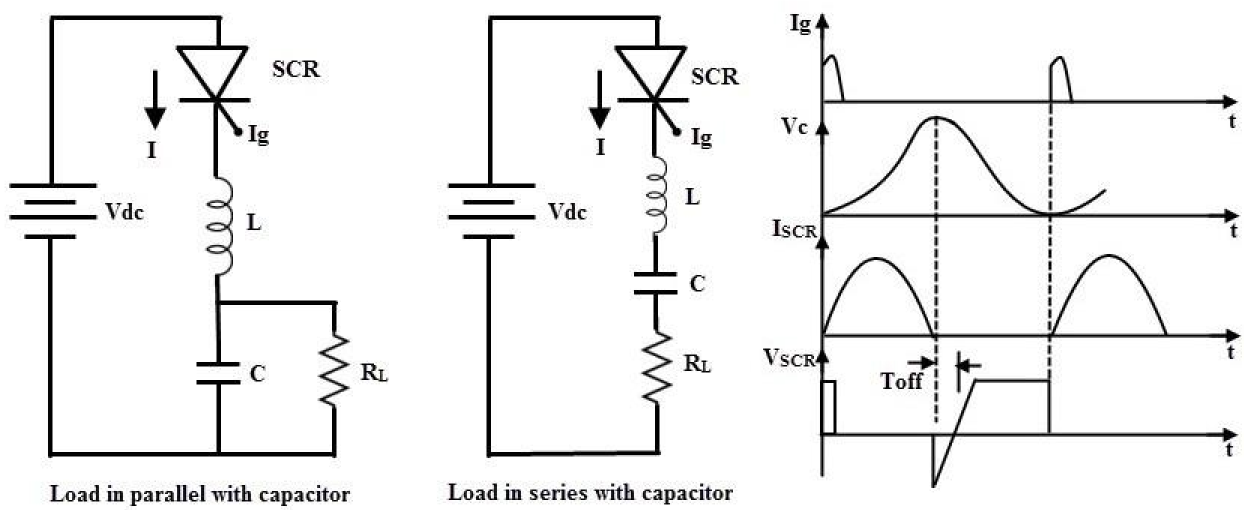
This method of commutation is also called as source commutation, or line commutation, or class F commutation. This commutation is possible with line commutated inverters, controlled rectifiers, cycloconverters and AC voltage regulators because the supply is the AC source in all these converters.



LOAD COMMUTATION CLASS A COMMUTATION

This is also known as self commutation, or resonant commutation, or load commutation. In this commutation, the source of commutation voltage is in the load. This load must be an under damped R-L-C supplied with a DC supply so that natural zero is obtained.

The commutating components L and C are connected either parallel or series with the load resistance R as shown below with waveforms of SCR current, voltage and capacitor voltage.



The value of load resistance and commutating components are so selected that they forms a under damped resonant circuit to produce natural zero. When the thyristor or SCR is triggered, the forward currents starts flowing through it and during this the capacitor is charged up to the value of E.

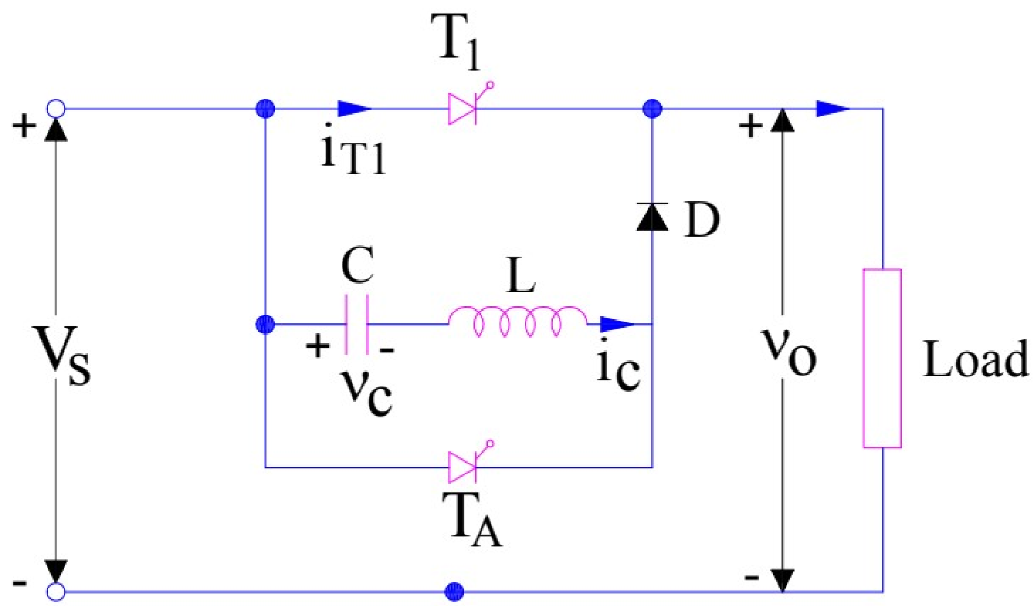
Once the capacitor is fully charged (more than the supply source voltage) the SCR becomes reverse biased and hence the commutation of the device. The capacitor discharges through the load resistance to make ready the circuit for the next cycle of operation. The time for switching OFF the SCR depends on the resonant frequency which further depends on the L and C components.

This method is simple and reliable. For high frequency operation which is in the range above 1000 Hz, this type of commutation circuits is preferred due to the high values of L and C components.

RESONANT PULSE COMMUTATION

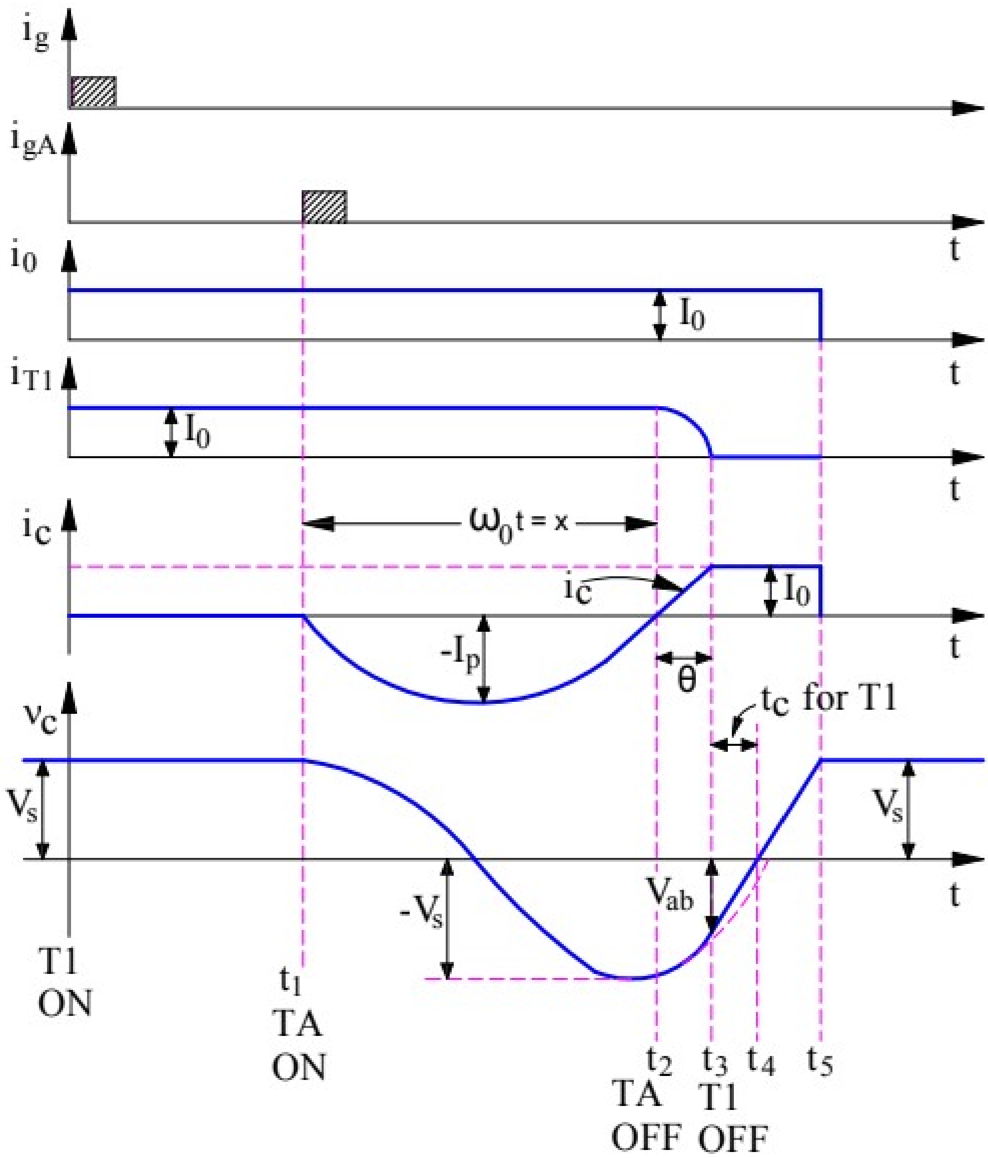
CLASS-B OR RESONANT PULSE COMMUTATION is a forced commutation technique to turn off an SCR. In this technique, thyristor or SCR is turned off by gradual build-up of resonant current in the reverse direction i.e. from cathode to anode of SCR. This technique is also known as current commutation and occurs in DC circuit not in AC circuit.

Let us consider the circuit diagram for Class-B or Resonant Pulse Commutation for better understanding of the commutation process involved.



The commutation circuit comprises of Capacitor C, Inductor L and an auxiliary thyristor TA. Initially thyristor T_1 and TA are in off state and capacitor C is charged to voltage V_s with left hand plate positive
as shown in figure. Positive direction of capacitor voltage and capacitor current i_c are shown in figure and taken as reference.

Now, at t=0, the main thyristor / SCR is gated and turned on. Load current equal to I_0 starts flowing through the main thyristor T_1 and Load. Now, we want to turn this thyristor off. To do this, we fire the auxiliary thyristor TA at $t=t_1$. Till time $t=t_1$, the capacitor is charged with source voltage V_s i.e. $v_c = V_s$, capacitor current $i_c = 0$ and current through main thyristor T_1 i.e. $i_0 = I_0$. This is shown in figure below.



When auxiliary thyristor TA is fired, it starts conducting and provides a path for the discharge of capacitor C. L, C and TA forms a resonating circuit. The resonating current i_c for this circuit is given as

$$ic = -Vs\sqrt{(\frac{C}{L})}Sin\omega t$$

$$ic = -ImSin\omega t$$

Negative sign in the above expression of resonating current is given as the actual current flows in a direction opposite to the direction of current i_c shown in the first figure.
Carefully observe the waveform of i_c. It can be seen that, after half cycle, the value of i_c reduces to zero at $t=t_2$. This means, the current through the auxiliary thuristor TA reduces to zero. Let's check if the auxiliary thyristor gets reversed biased after $t=t_2$. *Why are we checking this?* This is

because, the current through TA is zero at $t=t_2$ and if it gets reversed biased after $t=t_2$ then TA will get turned off. The voltage across TA equals the voltage across capacitor. The expression for capacitor voltage can be calculated as

$$vc = (1/C)\int icdt$$

$$= VsCos\omega t$$

$$Where\ \omega = Resonant\ Frequency$$

$$= 1/\sqrt{LC}$$

From the above expression of voltage across capacitor, if we put $\omega t = \pi$ then value of $Cos\omega t$ will -1. This means, the capacitor voltage will get reversed after half a cycle of capacitor current i.e. at $t=t_2$. Thus, the auxiliary thyristor TA is reversed biased after $t=t_2$. Hence it will get turned off at $t=t_2$.

Now, TA is OFF and capacitor C is charged up to source voltage Vs with its right hand plate positive. This means, the diode D is now forward biased and hence resonating current i_c will now flow through least resistive path i.e. through C, L, D and main thyristor T_1. But this resonating current ic will flow through the main SCR T_1 from cathode to anode i.e. in reverse direction. This simply means, the current I through the main thyristor T_1 will be given as

$I = I_0 - i_c$

When the magnitude of i_c reaches I_0, the current through the SCR T_1 will become zero. This can be seen at $t=t_3$. Now, you might ask, when the resonating current will attain a value of I_0? This can easily be calculated from the equation of the resonating current. Let's find it.

$$Vs\sqrt{\left(\frac{C}{L}\right)}Sin\omega(t3 - t2) = I0$$

$$Sin\omega(t3 - t2) = \left(\frac{I0}{Im}\right)$$

$$\omega(t3 - t2) = Sin^{-1}\left(\frac{I0}{Im}\right)$$

$$Where\ \omega = \frac{1}{\sqrt{LC}}\ and\ Im = Vs\sqrt{\left(\frac{C}{L}\right)}$$

Now, at $t=t_3$, the current through the main thyrsistor T_1 is zero. Let's check, if it is reversed biased at this instant of time. Again, the voltage across the main SCR T1 at this instant of time ($t=t_3$) is equal to the capacitor voltage. The capacitor voltage after $\omega t = \pi$ is negative. This means, the right hand plate is positive whereas left hand plate is negative. Hence, the main thyristor T_1 is reversed biased. Thus, main thyristor T_1 will turn off at $t=t_3$ as the current through it is zero and it is reversed biased after this instant of time.

From the above discussion, it should have been clear that the peak value of resonating current i_c i.e. I_m in the expression of i_c, must be more than load current (I_0) for reliable commutation of thyristor / SCR. As SCR is commutated by the gradual build-up of the resonating current ic in the reverse direction of SCR, this method of commutation is called the *current commutation, resonant pulse commutation or Class-B commutation*.

Let's now check what happens after the commutation of main SCR T_1. Once the main SCR T_1 is turned off, load current I0 begins to flow from source V_s to load though C, L and D. This causes capacitor C to charge linearly from V_{ab} to zero at $t = t_4$ and then to source voltage V_s at $t=t_5$. At $t=t_5$, the capacitor is charged up to source voltage V_s with its left hand plate positive. Therefore, capacitor will not allow the flow of load current after $t = t_5$.

The circuit turn off time is equal to the time period for which the main thyristor / SCR is reversed biased. Here, this time period is $(t_4 – t_3)$. Therefore,

Circuit Turn Off time t_c for Class-A commutation = $(t_4 – t_3) = (V_{ab}C) / I_0$.

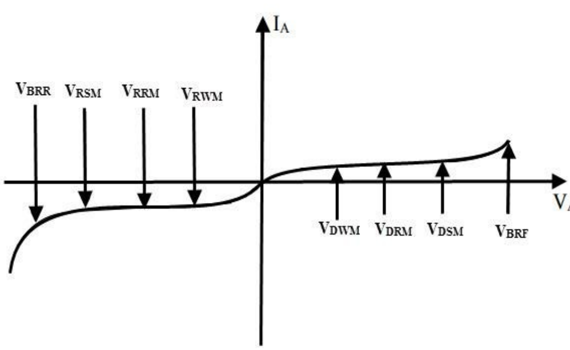
VOLTAGE AND CURRENT RATINGS OF SCR.

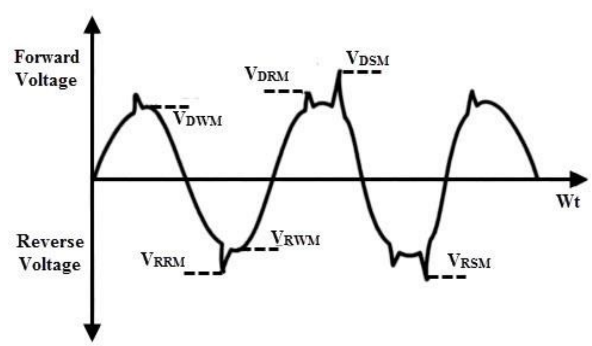
VOLTAGE RATINGS OF SCR

The voltage capability of the SCR should not be exceeded during the operation even for short periods. So the SCR is assigned with different voltage ratings, which are the maximum voltages at which the SCR can function normally without breakdown of junctions. These are assigned in both blocking states of an SCR and can withstand against voltage transients. The various voltage ratings of an SCR are given below.

PEAK WORKING FORWARD-BLOCKING VOLTAGE V_{DWM}

It specifies the maximum instantaneous value of forward blocking voltage across the SCR excluding all surge and repetitive transient voltages. Beyond this value of the voltage the SCR cannot withstand during its operation. This VDWM is equal to the maximum or peak value of the supply voltage wave shown in figure.





PEAK REPETITIVE FORWARD-BLOCKING VOLTAGE V_{DRM}

It is the maximum transient voltage that the SCR can block during it's the forward blocking state repeatedly or periodically. This is specified with a specific biasing resistance between cathode and gate or at a maximum permissible junction temperature with gate circuit open.

This voltage V_{DRM} is encountered or appeared across the SCR , when the SCR is turned OFF or commutated or due to diodes in the converter circuit. During the turn OFF process, an abrupt change in reverse recovery current causes to create a voltage spike , which is responsible of V_{DRM} to appear across the SCR.

PEAK NON-REPETITIVE OR SURGE FORWARD-BLOCKING VOLTAGE V_{DSM}

This is the maximum instantaneous value of forward surge voltage across the SCR that is of non-repetitive. This V_{DSM} is less than the forward break over voltage V_{BO} and this value is in the range about 130 percent of V_{DRM}.

PEAK WORKING REVERSE VOLTAGE V_{RWM}

This is the maximum instantaneous value of reverse voltage across the SCR excluding all surge and repetitive transient voltages. This V_{RWM} is equal to the maximum negative value of the supply voltage wave shown in figure.

PEAK REPETITIVE REVERSE VOLTAGE V_{RRM}

It is occurrence of the maximum reverse transient voltage repeatedly or periodically across the SCR in the reverse direction at permissible maximum junction temperature. Beyond this rating the SCR may get damaged due to excessive junction temperature. This voltage is also appeared due to the same reason as of V_{DRM}.

PEAK NON-REPETITIVE OR SURGE REVERSE VOLTAGE V_{RSM}

It refers to the maximum value of reverse transient voltage across the SCR that is of non-repetitive. This V_{RSM} is less than the reverse break over voltage V_{BR} and this value is in the range about 130 percent of V_{RRM}. The surge voltage ratings V_{DSM} and V_{RSM} can be increased by connecting a diode of equal current rating in series with the SCR.

The above discussed voltage ratings are belonging to the forward and reverse blocking states with which the SCR is able to withstand with gate open.

ON-STATE VOLTAGE V_T

This is the voltage drop between the anode and cathode with specified junction temperature and ON-state forward current. Generally, this value is in the order of 1 to 1.5 Volts.

GATE TRIGGERING VOLTAGE V_{GT}

This is the minimum voltage required by the gate to produce the gate trigger current.

FORWARD DV/DT RATING

This is the maximum rate of rise of anode voltage that will not trigger the SCR without any gate pulse or signal. If this value is more than the specified value, the SCR may be switched ON. The SCR in forward blocking mode is analogous to the capacitor with a dielectric.

So , the charging current flows through it when the applied voltage is increased. If the rate of rise of voltage is more, sufficient charges will flow through the junctions J_2 of the SCR and hence the SCR will be turned ON without any gate signal.

This type of triggering is called as false triggering and in practice it is not employed. Also, this rating depends on the junction temperature. If the junction temperature is high, the dv/dt rating of the SCR is lower and vice-versa. With the use of snubber networks across the SCRs, it is possible to limit the maximum dv/dt applied to the SCR.

VOLTAGE SAFETY FACTOR V_F

Generally, the operating voltage of the SCR is kept below the V_{RSM} to avoid the damage to the SCR due to uncertain conditions. Therefore, the voltage safety factor relates the operating voltage and V_{RSM} and is given as

$V_f = V_{RSM}$ / (RMS value of the input voltage * $\sqrt{2}$)

CURRENT RATINGS OF SCR

Basically an SCR is a unilateral device and hence average current rating is assigned to it (while RMS current rating is assigned to bilateral devices). An SCR has low thermal capacity and short time constant. This means the junction temperature exceeds its rated value even for short over current.

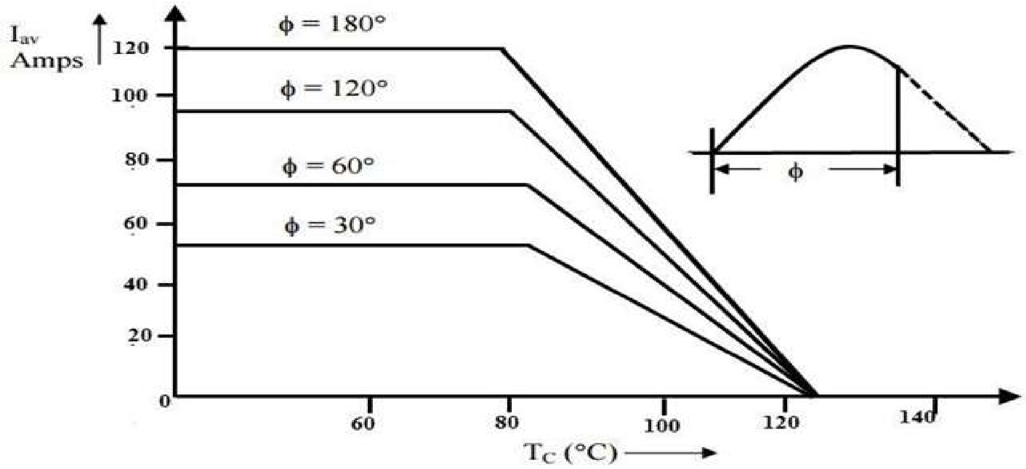
This may lead to damage the SCR. Therefore, current ratings must be properly selected for long life of SCR, as the junction temperature depends on the current handled by it. Let us look at various current ratings of an SCR.

AVERAGE ON-STATE CURRENT RATING I_{TAV}

This is the maximum repetitive average value of forward current that can flow through the SCR such that the maximum temperature and RMS current limits are not exceeded. The forward voltage drop across the SCR is very low when it is in conduction mode. So the power loss in the thyristor is entirely depends on the forward current I_{TAV}.

In case of phase controlled SCRs, average forward current depends on the firing angle. For the given average forward current, the RMS value of the current is increased with decrease in conduction angle. This leads to increase the voltage drop across the SCR which in turn increases the average power dissipation. Hence the junction temperature rises beyond the safe limit.

In order to limit the maximum junction temperature, the permissible average forward current has to be lowered with decrease in conduction angle. The manufacturers usually provide the data sheet that shows the forward average current variation with respect to the case temperature. As an example the current waveform formed from the positive half cycle for different conduction angles is shown in below.



RMS ON-STATE CURRENT I_{TRMS}

This is the maximum repetitive RMS current specified at a maximum junction temperature that can flow through the SCR. For a direct current, both RMS and average currents are same. However, this rating is important for SCRs subject to low duty waveforms with peak currents. And also this rating is required to prevent excessive heating in leads, metallic joints and interfaces of SCR.

SURGE CURRENT RATING I_{TSM}

It specifies the maximum non-repetitive or surge current that the SCR can withstand for a limited number of times during its life span. The manufacturers specify the surge rating to accommodate the abnormal conditions of SCR due to short circuits and faults. If the peak amplitude and the number of cycles of the surge current are exceeded, the SCR may get damaged.

I^2_T RATING

This rating is used to determine the thermal energy absorption of the device. This rating is required in the choice of a fuse or other protective equipment employed for the SCR. This is the measure of the thermal energy that the SCR can absorb for a short period of time before clearing the fault by the fuse.

It is the time integral of the square of the maximum instantaneous current. For a reliable protection of SCR by the fuse or other protective equipment, the I^2_t rating of the fuse (or any other protective equipment) must be less than the I^2_t rating of the SCR.

DI/DT RATING

It is the maximum allowable rate of rise of anode to cathode current without any damage or harm to an SCR. If the rate of rise of anode current is very rapid compared to the spreading velocity of the charge carriers, local hot spots are created due to concentration of carriers (on account of high current density) in the restricted area of the junctions.

This raises the junction temperature above the safe limit and hence the SCR may be damaged. Therefore, for all SCRs the maximum allowable di/dt rating specified in order to protect the SCR. It is specified in amperes/microseconds and typically it lies in the range 50 to 800 ampere/microseconds.

LATCHING CURRENT I_L

It is the minimum ON state current required to maintain the SCR in ON state after gate drive has been removed. After turning ON of the SCR, the anode current must be allowed to build up such that the latching current is attained before the gate pulse is removed. Otherwise the SCR will be turned OFF if the gate signal is removed.

HOLDING CURRENT I_H

This is the minimum value of the anode current below which SCR stops conducting and turns OFF. The holding current is associated with turn OFF process and usually it is a very small value in the range of mill amperes.

GATE CURRENT I_G

As the gate current is more, earlier will be the turn ON of the SCR and vice-versa. However, safety limits must be provided for gate by specifying maximum and minimum gate currents. For controlling the SCR, gate current is applied to the gate terminal. This gate current is divided into two types; minimum gate current I_{Gmin} and maximum gate current I_{Gmax}.

The minimum gate current I_{Gmin} is the current required by the gate terminal to turn ON the SCR where as I_{Gmax} is the maximum current that can be applied safely to the gate. Between these two limits the conduction angle of the SCR is controlled.

TEMPERATURE RATING OF SCR

The forward and reverse blocking capability of the SCR is determined by junction temperature T_j. If the maximum junction temperature is exceeded, the SCR will be driven to conduction state even without any gate signal. This upper limit of T_j is imposed by considering the temperature dependence on break over voltage, thermal stability and turn OFF time.

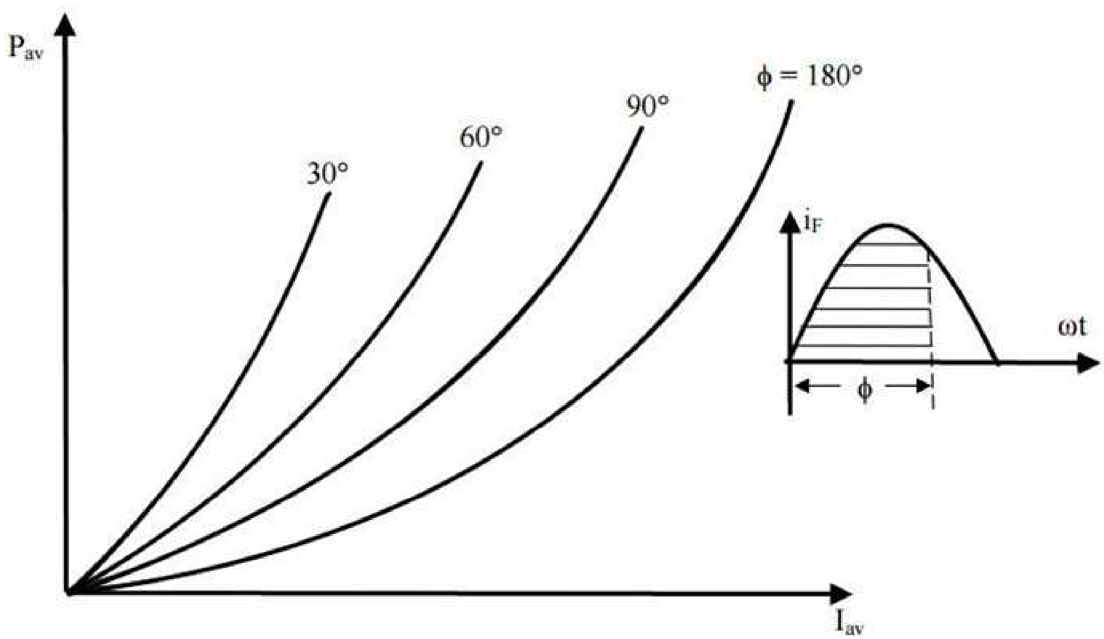
And also an upper storage temperature limit T_s is also required to limit thermal stresses on silicon crystal, lead attachments and encapsulating epoxy. Excess of these two temperature limits may cause unreliable operation of an SCR. In some cases, upper storage temperature limit is higher than the operating temperature limit of an SCR.

POWER RATINGS OF SCR

The power dissipation in the SCR produces a temperature rise in the junction regions. The dissipation of power in the SCR includes forward power dissipation; turn ON and OFF losses and gate power dissipation.

AVERAGE POWER DISSIPATION P_{AV}

It is the multiplication of the average anode current and forward voltage drop across the SCR. This is the major source of junction heating in an SCR for normal duty cycle operations. The peak power from a given source must not exceed the average power dissipation rating to maintain the safety of the device. This rating is specified for different conduction angles as a function of average forward current as shown in figure.



GATE POWER DISSIPATION P_G

This rating defines both forward or reverse peak power and the average power applied to the gate. If these ratings are exceeded, considerable damage occurs to the gate. Therefore, while calculating the voltage and currents applied, the width of gate pulses has to be considered (because the peak power is the function of time). For pulse type triggering, gate losses are negligible whereas gate signals with a high duty cycle, the gate losses becomes more significant.

Other power losses include ON state losses, OFF state losses, forward blocking losses and reverse blocking losses. Turn ON and OFF losses have to be taken into consideration while selecting the SCR rating since these constitute a significant portion of the total losses. And also forward and reverse blocking losses are very small compared to the conduction losses since a small leakage current and negligible voltage drop in blocking states.

TURN ON AND TURN OFF TIME RATINGS

The turn ON time is the time interval between the instant at which the gate signal is applied and the instant at which the ON-state current reaches 90 percent of its final value. Shorter will be the turn ON time if the gate drive is increased. This turn ON time is valid only for resistive load because the rate of rise of anode current is slow in inductive load.

Therefore, the turn ON time does not indicate the time in which the device stays ON if the gate signal is removed. And if the load is resistive, turn ON time surely, indicates the time interval in which the SCR stays ON even the gate is removed.

Turn OFF time is the time interval between the instant at which the anode current goes zero or negative and the instant positive voltage is reapplied to the SCR. For fast switching SCRs both turn ON and OFF time values are very low.

PROTECTION OF SCR

Thyristors are very sensitive to Overvoltages just as other Semiconductor devices are. Overvoltage transients are perhaps the main cause of thyristor failure. Transient Overvoltages cause either maloperation of the circuit by unwanted turn on of a thyristor or permanent damage to the device due to reverse breakdown. A thyristor may be subjected to internal or external Overvoltages; the former is caused by the thyristor operation where as the letter comes from the supply lines or the load circuit.

OVERVOLTAGE PROTECTION

Internal Overvoltages

Internal over voltages arise while the SCR is in operation. During the turn OFF of an SCR, a reverse current continues to flow through the SCR after the anode current decreased to zero to sweep away the earlier stored charge. This reverse current decay at a faster rate at the end of reverse recover interval.

Due to the inductance of the circuit, this high di/dt produces a high voltage. This voltage value may be much higher than the rated value of the SCR and hence the SCR may be damaged.

External Over Voltage

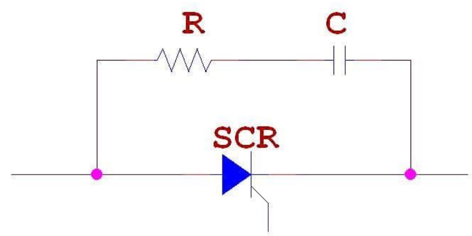
This is due to external supply and load condition. This is because of the interruption of current flow in an inductive circuit and also due to lightening strokes on the lines feeding the thyristor systems. When a SCR converter is fed from a transformer, voltage transient occur when transformer primary will energize or de-energised.

Search Overvoltages may cause random turn on of a thyristor as a result the Overvoltages may appear across the load causing the flow of large fault currents Overvoltages may also damage the thyristor by an inverse breakdown. For reliable operation the Overvoltages must be suppressed by adopting suitable techniques.

In order to keep the protective components to a minimum, thyristors are chosen with their peak voltage ratings of 2.5 to 3 times their normal peak working voltage. The effect of Overvoltages is usually minimized by using RC circuits and non-linear registers called voltage clamping devices.

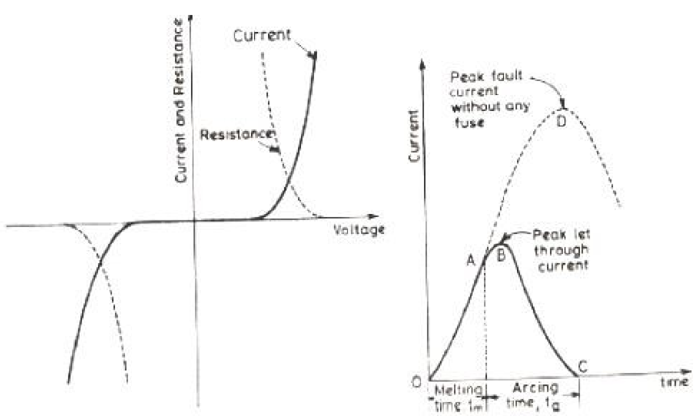
Snubber circuit

The RC circuit called snubber circuit is connected across the device to be protected as shown in fig. It provides a local path for internal overvoltage caused by reverse recovery current .Snubber circuit is also helpful in damping overvoltage transient spikes and for limiting dv/dt across the thyristor. The capacitor charges at a slow rate and thus the rate of rise of Forward voltage dv/dt across SCR is also reduced. The resistance R_S damps out the ringing oscillations between the Snubber circuit and the stray circuit inductance.



Voltage clamping device

It is a non-linear resistor called as VARISTOR (VARIable resiSTOR) connected across the SCR. The resistance of varistor will decrease with increase in voltage. During normal operation, varistor has high Resistance and draws only small leakage current. When high voltage appears, it operates in low resistance region and the surge energy is dissipated across the resistance by producing a virtual short-circuit across the SCR.



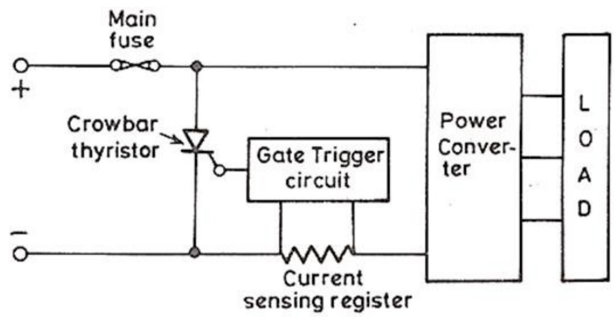
OVER CURRENT PROTECTION

OVER CURRENT PROTECTION:

In an SCR due to over-current, the junction temperature exceeds the rated value and the device gets damaged. Over-current is interrupted by conventional fuses and circuit breakers. The fault current must be interrupted before the SCR gets damaged and only the faulty branches of the network should be isolated. Circuit breaker has long tripping time. So it is used for protecting SCR against continuous over loads (or) against surge currents of long duration. Fast acting current limiting fuse is used to protect SCR against large surge currents of very short duration.

Electronic Crowbar Protection:

SCR has high surge current ability.SCR is used in electronic crowbar circuit for overcurrent protection of power converter. In this protection, an additional SCR is connected across the supply which is known as 'Crowbar SCR'. Current sensing resistor detects the value of converter current. If it exceeds preset value, then gate trigger circuits turn ON the crowbar SCR. So the input terminals are short-circuit by SCR and thus it bypass the converter over current. After some time the main fuse interrupts the fault current.

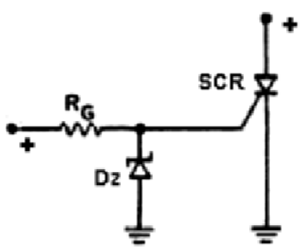


GATE PROTECTION

The first circuit adds resistor Rg to limit the current into the gate of the SCR and a Zener diode to protect against voltage transients.

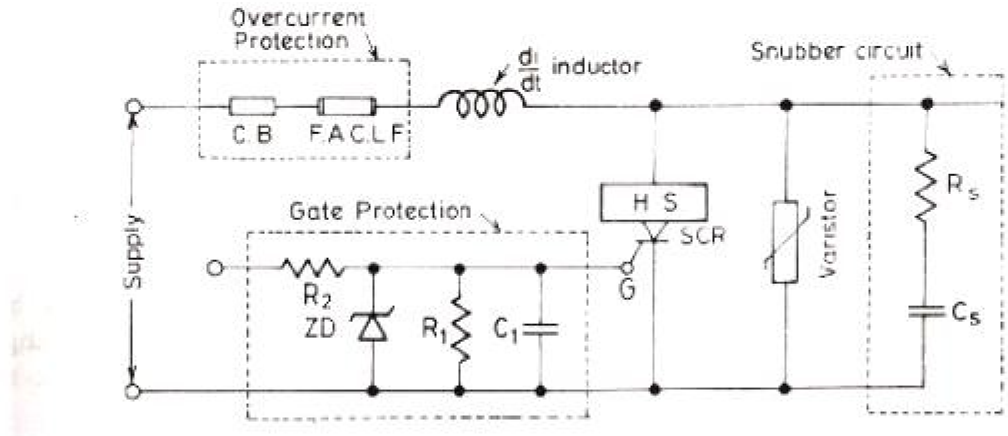
When the gate of the SCR is turned on the Zener diode [Dz] protects against any voltage transients.

A capacitor and resister is connected across gate to cathode to bypass noise signal.



SCR Protection Circuit

The diagram below shows the various components used to provide overall protection to scr.

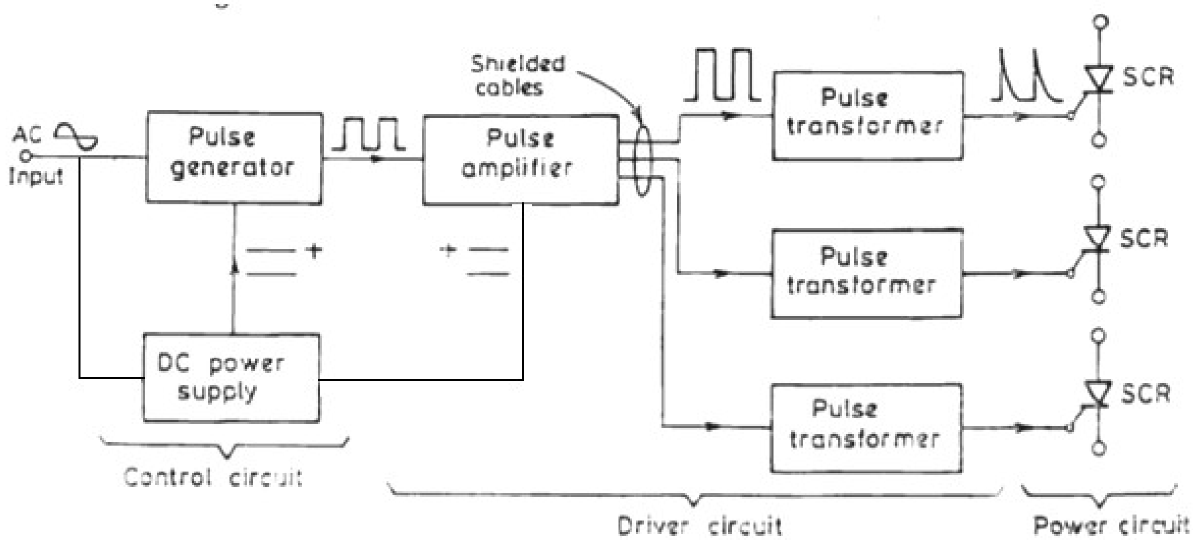


Thyrister protection circuit

FIRING CIRCUITS

GENERAL LAYOUT DIAGRAM OF FIRING CIRCUIT

The most common method for controlling the onset of conduction in an SCR is by means of gate voltage control. The gate control circuit is also called firing, or triggering, circuit. These gating circuits are usually low-power electronic circuits. A firing circuit should fulfill the following two functions.



A geneal layout of firing circuit scheme for scr

(i) If power circuit has more than one SCR, the firing circuit should produce gating pulses for each SCR at the desired instant for proper operation of the power circuit. These pulses must be periodic in nature and the sequence of firing must correspond with the type of thyristorised power controller. For example, in a single-phase semiconverter using two SCRs, the triggering circuit must produce one firing pulse in each half cycle; in a 3-phase full converter using SCRs, gating circuit must produce one trigger pulse after every 60° interval.

(ii) The control signal generated by a firing circuit may not be able to turn-on an SCR It is therefore common to feed the voltage pulses to a driver circuit and then to gate-cathode circuit. A driver circuit consists of a pulse amplifier and a pulse transformer. A firing circuit scheme, in general, consists of the components shown in Fig. A regulated dc power supply is obtained from an alternating voltage source. Pulse generator, supplied from both ac and dc sources, gives out voltage pulses which are then fed to pulse amplifier for their amplification. Shielded cables transmit the amplified pulses to pulse transformers. The function of pulse transformer is to isolate the low-voltage gate-cathode circuit from the' high-voltage anode-cathode circuit. Some firing circuit schemes are described in this section.

DRAW R FIRING CIRCUITS AND EXPLAIN

The circuit below shows the resistance triggering of SCR where it is employed to drive the load from the input AC supply. Resistance and diode combination circuit acts as a gate control circuitry to switch the SCR in the desired condition.

As the positive voltage applied, the SCR is forward biased and doesn’t conduct until its gate current is more than minimum gate current of the SCR.

When the gate current is applied by varying the resistance R2 such that the gate current should be more than the minimum value of gate current, the SCR is turned ON. And hence the load current starts flowing through the SCR.

The SCR remains ON until the anode current is equal to the holding current of the SCR. And it will switch OFF when the voltage applied is zero. So the load current is zero as the SCR acts as open switch.

The diode protects the gate drive circuit from reverse gate voltage during the negative half cycle of the input. And Resistance R1 limits the current flowing through the gate terminal and its value is such that the gate current should not exceed the maximum gate current.

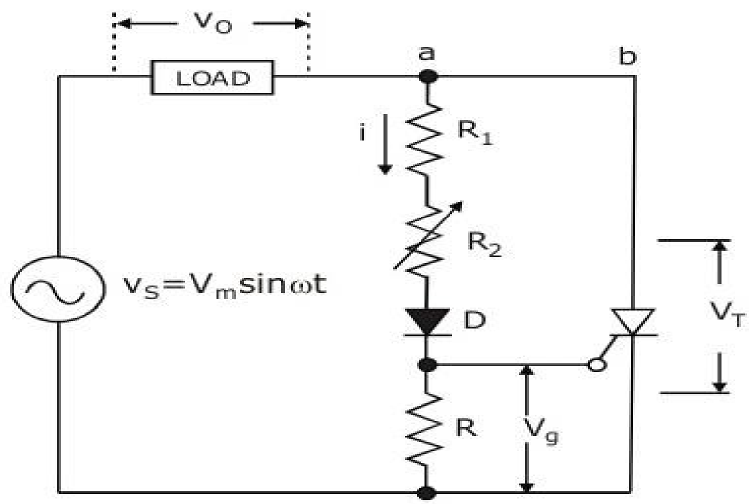
$$\frac{v_m}{R_1 + R} \times R \leq V_{gm}$$

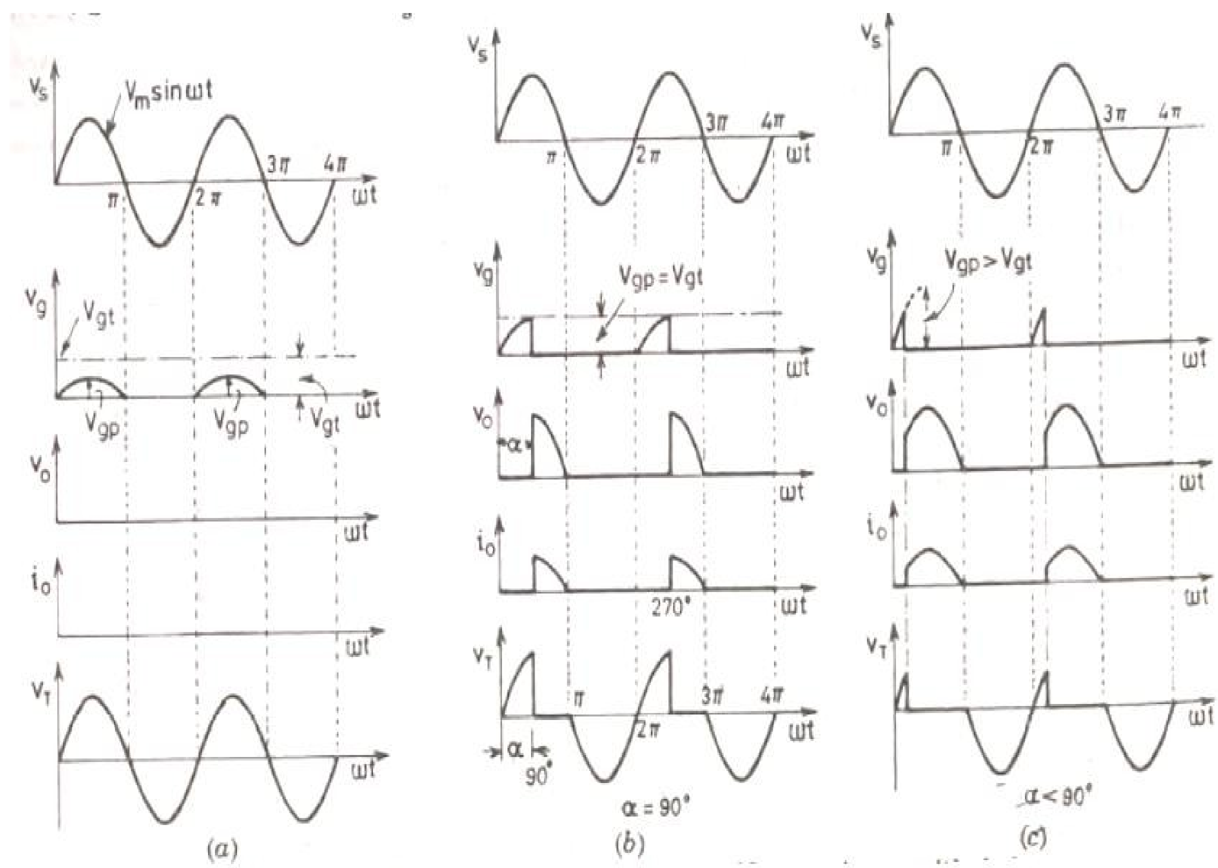
$$R \leq \frac{V_{gm} \cdot R_1}{v_m - V_{gm}}$$

It is the simplest and economical type of triggering but limited for few applications due to its disadvantages.

In this, the triggering angle is limited to 90 degrees only. Because the applied voltage is maximum at 90 degrees so the gate current has to reach minimum gate current value somewhere between zero to 90 degrees.

R triggering circuit with its waveform





Resistance firing of an scr in a half wave circuit with dc load
(a)No triggering of scr (b)α=90° (c) α<90°

RESISTANCE – CAPACITANCE (RC) FIRING CIRCUIT

The limitation of resistance firing circuit can be overcome by the RC triggering circuit which provides the firing angle control from 0 to 180 degrees. By changing the phase and amplitude of the gate current, a large variation of firing angle is obtained using this circuit.

Below figure shows the RC triggering circuit consisting of two diodes with an RC network connected to turn the SCR.

By varying the variable resistance, triggering or firing angle is controlled in a full positive half cycle of the input signal.

During the negative half cycle of the input signal, capacitor charges with lower plate positive through diode D2 up to the maximum supply voltage Vmax at $\omega t=90^\circ$.

After $\omega t=90^\circ$ source voltage v_s decreases from $-v_m$ at $\omega t=90^\circ$ to 0 at $\omega t=0^\circ$.During this period capacitor voltage v_c may fall from $-v_m$ at $\omega t=90^\circ$ to some lower value –oa at $\omega t=0^\circ$.

During the positive half cycle of the input, the SCR becomes forward biased and the capacitor starts charging through variable resistance R to the triggering voltage value of the SCR.

When the capacitor charging voltage is equal to the gate trigger voltage, SCR is turned ON and the capacitor holds a small voltage. Therefore the capacitor voltage is helpful for triggering the SCR even after 90 degrees of the input waveform.

In this, diode D1 prevents the negative voltage between the gate and cathode during the negative half cycle of the input through diode D2.

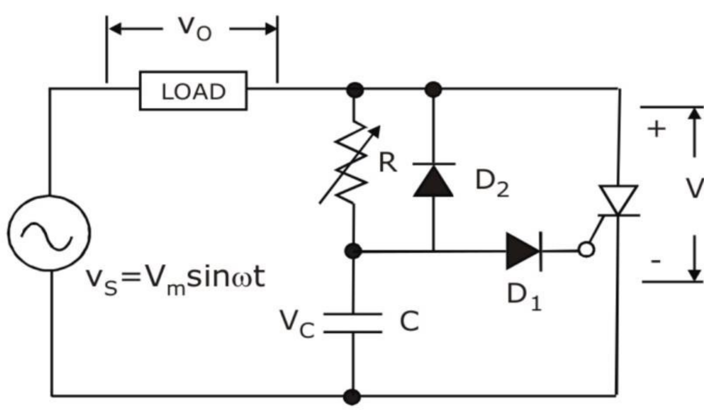
The value of rc is given by $RC \geq \frac{1.3T}{2} \cong \frac{4}{\omega}$ where T=1/f

The SCR will trigger when $v_c=v_{gt}+v_d$.At the instant of triggering, if v_c is assumed constant,the current i_{gt} must be supplied by voltage source through R,D_1 and gate to cathode circuit hence maximum value of R is given by

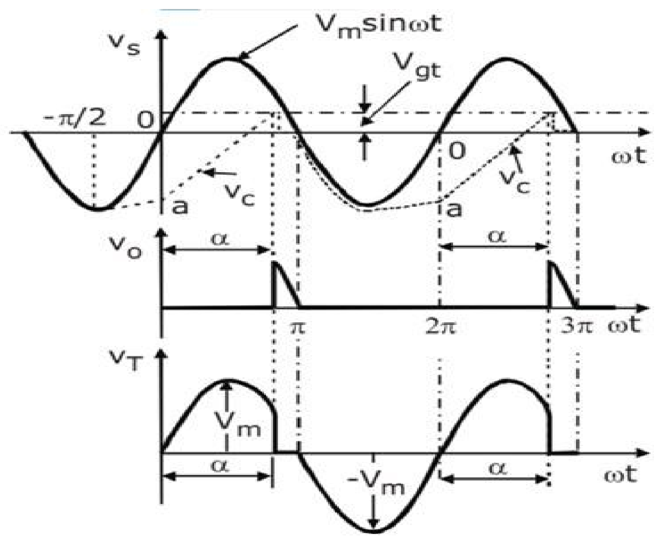
$v_s \geq RI_{gt} + V_c$

$v_s \geq RI_{gt} + v_{gt} + v_d$

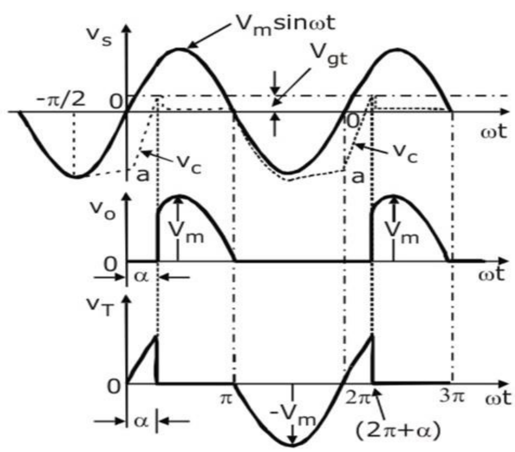
$$R \leq \frac{vs - vgt - vd}{Igt}$$



RESISTANCE – CAPACITANCE (RC) FIRING CIRCUIT



High Value of R



Low Value of R

RC FULL WAVE TRIGGER CIRCUIT

Power can be delivered to the load in Half wave during the positive half-cycle of v_s because the SCR conducts only when it is forward biased. This limitation can be overcome in several ways, one of which is shown in Fig. below Here; the ac line voltage is converted to pulsating dc. by the full-wave diode bridge. This allows the SCR to be triggered "on" for both half-cycle of the line voltage, which doubles the available power to the load.

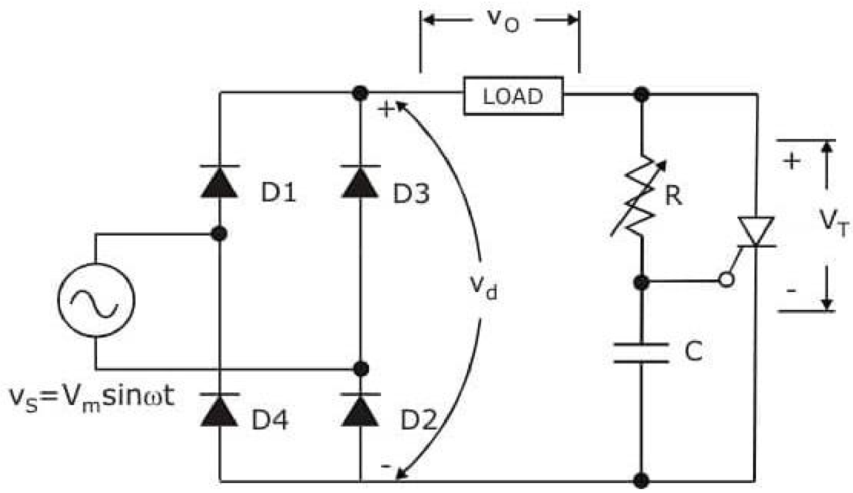
In this circuit, the initial voltage from which capacitor *C* charges is almost zero. Capacitor *C* is set to this low positive voltage (upper plate positive) by the clamping action of SCR gate. When the capacitor charges to a voltage equal to Vgt SCR triggers and rectified voltage Vdc appears across load as Vo. The value of RC is obtained from the following relation:

$$RC \geq 50\frac{T}{2} \cong \frac{157}{\omega}$$

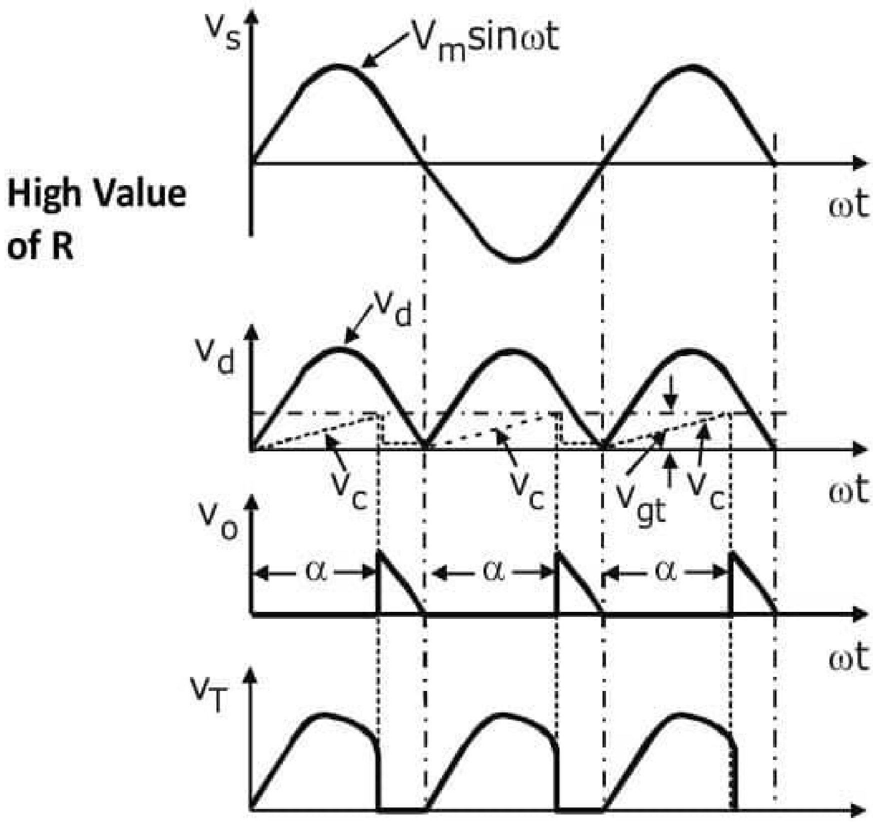
The value of R is given by

$$R \ll \frac{Vs - Vgt}{Igt}$$

Where Vs is the source voltage at which thyrister turns on.in this fig firing angle α is more than 90˚.

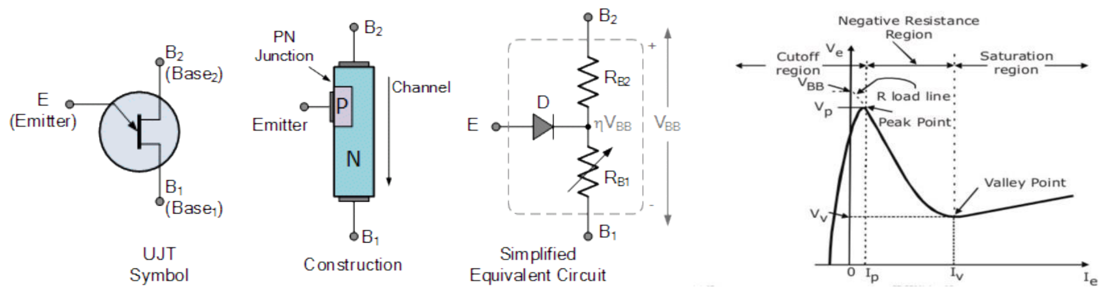


RC full wave trigger circuit diagram



RC full wave trigger waveform

UJT PULSE TRIGGER CIRCUIT



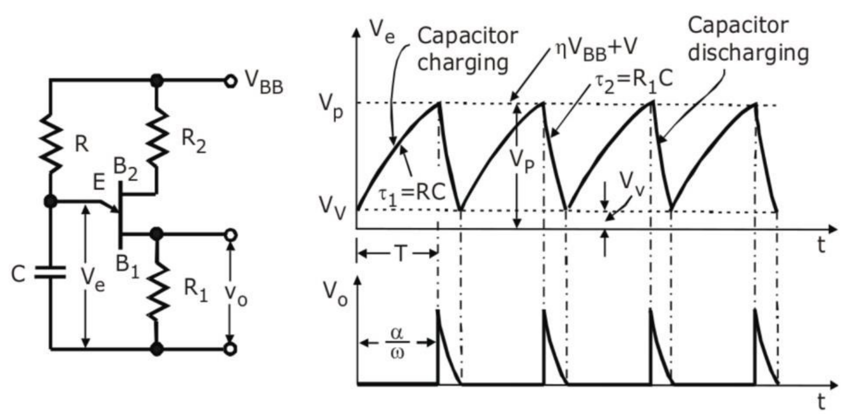
VI characterstics of ujt

The unijunction transistor is a highly efficient switch; its switching time is in the range of nanoseconds. Since UJT exhibits negative resistance characteristics, it can be used as a relaxation oscillator.
The Fig. below shows a circuit diagram with UJT working in the oscillator mode. The external resistances R 1.R2 are small in comparison with the internal resistances RB1, RB2 of UJT bases. The charging resistance R should be such that its load line intersects the device characteristics only in the negative resistance region.
In this Fig. when source voltage VBB is applied. capacitor C begins to charge through R exponentially towards VBB' During this charging, emitter circuit of UJT is an open circuit. The capacitor voltage Vc equal to emitter voltage Ve is given by

$Vc=Ve=V_{BB}(1-e^{-t/rc})$

The time constant of the charge circuit is τ_1=RC

The time constant of the charge circuit is τ_1=RC. When this emitter voltage Vc (or Ve) reaches the peak-point voltage Vp (= ηV_{BB} + VD)' the unijunction between E - BI breaks down. As a result, UJT turns on and capacitor C rapidly discharges through low resistance R1 with a time constant $\tau_2 = R_tC$. Here τ_2 is much smaller than τ_1. When the emitter voltage decays to the valley-point voltage Vv, emitter current (Vv/(R B1 + R1) falls below Iv and UJT turns off. The time T required for capacitor C to charge from initial voltage Vv to peak-point voltage Vp through large resistance R, can be obtained as under:



$$V_p = \eta V_{BB} + V_D = V_v + V_{BB}(1 - e^{-T/RC})$$

Assuming $$V_D = V_v, \quad \eta = (1 - e^{-T/RC})$$

or $$T = \frac{1}{f} = RC \ \ln\left(\frac{1}{1-\eta}\right)$$

In case T is taken as the time period of output pulse duration (neglecting small discharge time), then the value of firing angle α1 is given by $\alpha_1 = \omega T = \omega RC \ln(1/1-\eta)$

EXPLAIN SYNCHRONOUS TRIGGERING (RAMP TRIGGERING)

A synchronized UJT trigger circuit using an UJT is shown in Fig. Diodes Dl - D4 rectify ac to dc. Resistor Rl lowers Vdc to a suitable value for the zener diode and UJT. Zener diode Z functions to clip the rectified voltage to a standard level V_s which remains constant except near the Vdc zero as in Fig. This voltage V_s is applied to the charging circuit RC. Current i_1 charges a C at a rate determined by R. Voltage across capacitor is marked by C in Figs. when voltage V_s reaches the unijunction threshold voltage ηV_s, the E – B_1 junction of UJT breaks down and the capacitor C discharges through primary of pulse transformer sending a current i_2 as shown in Fig. As the current i2 is in the form of pulse, windings of the pulse transformer have pulse voltages at their secondary terminals. Pulses at the two secondary windings feed the same in-phase pulse to two SCRs of a full-wave circuit. SCR with positive anode voltage would turn on. As soon as the capacitor discharges, it starts to recharge as shown. Rate of rise of capacitor voltage can be controlled by varying R. The firing angle can be controlled up to about 150°C.

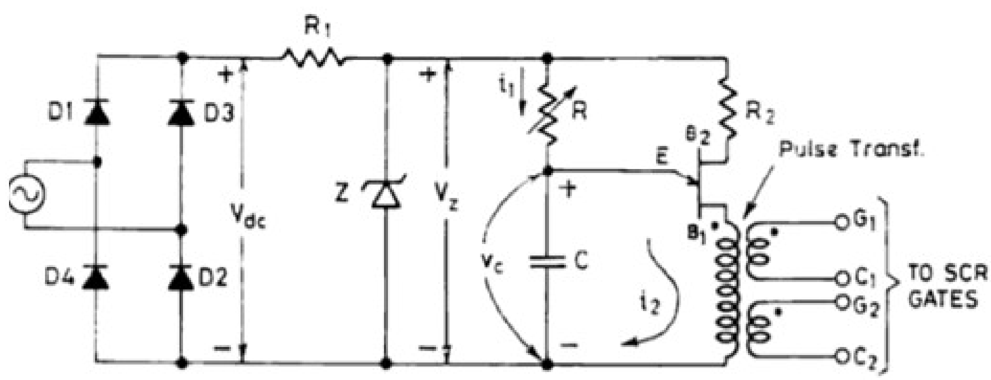


Fig. Synchronised UJT trigger circuit.

This method of controlling the output power by varying charging resistor R is called ramp control, open loop control or manual control. As the zener diode voltage Vz goes to zero at the end of each half cycle, the synchronization of the trigger circuit with the supply voltage across SCRs is achieved. Thus the time t, equal to α/ω, when the pulse is applied to SCR for the first time, will remain constant for the same value of R. Small variations in the supply voltage and frequency are not going to effect the circuit operation.

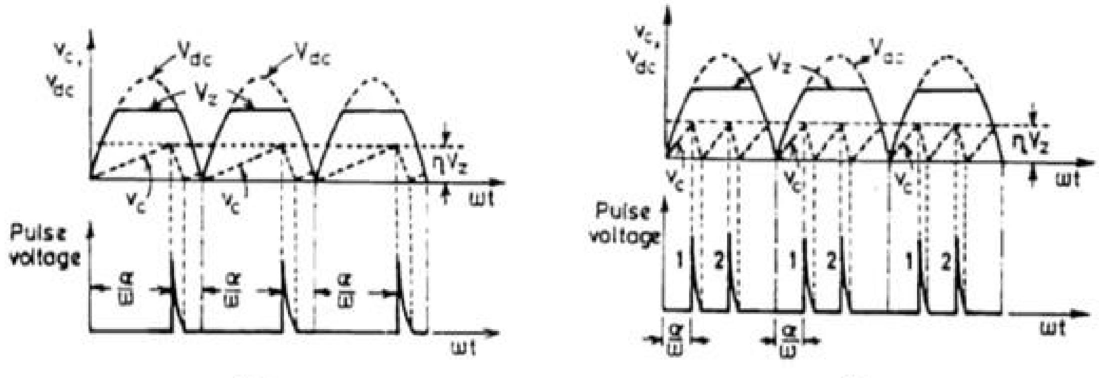


Fig. Generation of output pulses for the circuit Here, $t = \alpha/\omega$.

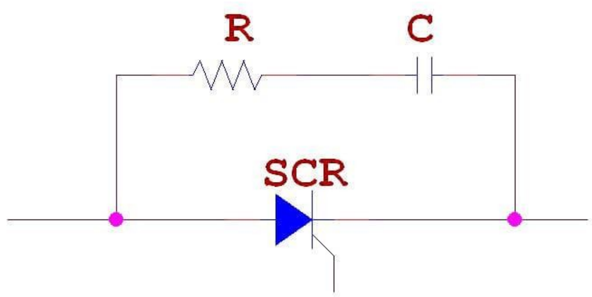
In case R is reduced so that Vc reaches UJT threshold voltage twice in each half cycle as shown in Fig. then there will be two pulses in each half cycle. As the first pulse will be able to turn-on the SCR, second pulse in each cycle is redundant.

DESIGN OF SNUBBER CIRCUITS

Power semiconductors are the heart of power electronics equipment. Snubbers are circuits which are placed across semiconductor devices for protection and to improve performance. Snubbers can do many things: ·

Reduce or eliminate voltage or current spikes · Limit dI/dt or dV/dt · Shape the load line to keep it within the safe operating area (SOA) · Transfer power dissipation from the switch to a resistor or a useful load · Reduce total losses due to switching · Reduce EMI by damping voltage and current ringing

With the help of snubber circuit, the false turn-on of a thyristor due to large dv/dt can be prevented.



RC Snubber Circuit for SCR dv/dt Protection:

This type of snubber circuit consists of a series combination of resistance R and Capacitance C in parallel with a SCR.

- When a reverse voltage is applied, commutation process is initiated and the forward current flow through SCR approaches zero.
- Due to the inductance, current continuous to flow due to sweeping of charge carries at the external junctions.
- When it reaches a peak value it cannot be further supported by the charge carriers and falls very quickly to zero. This causes a voltage spike with the value of L(di/dt).

- Also when the supply is closed to the circuit (in the above figure say the switch S is closed), sudden voltage appears across SCR.
- Now, as the thyristor current is zero it can be considered as an open switch.
- At this moment, the capacitor C behaves like a short-circuit and therefore voltage across the SCR is zero.
- With the passage of time capacitor C gets charged at a slow rate such that dv/dt across the capacitor and therefore across SCR is less than the specified maximum dv/dt rating of the device.
- Thus the capacitor protects the SCR against high voltages and high dv/dt.

Based on the above discussion we can say that simply a Capacitor C is sufficient to protect the SCR against dv/dt false triggering.

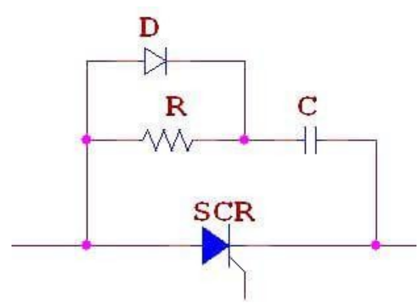
Then what is the purpose of resistance R?

- In the RC snubber circuit, the resistance R limits the discharge current of capacitor at the instant of firing of SCR.
- Before SCR is fired, capacitor C charges to full voltage V.
- If SCR is fired, when the capacitor voltage is maximum, it discharges through the local path formed by capacitor C, Resistance R and SCR.
- During this time, if the resistance R is not included in the circuit, the discharge current will be high and consequently may damage the SCR due to large di/dt.
- Thus the Resistance R in the snubber circuit reduces the discharge current of the capacitor C and thus protect the SCR against large di/dt.

In actual practice, R, C and the load current parameters should be such that

1. dv/dt across C during its charging is less than the specified dv/dt rating of the SCR
2. Discharge current at the turn ON of the SCR is within reasonable limits.

Normally R,S and load circuit parameters form an under damped circuit so that dv/dt is limited to acceptable values.



In some RC snubber circuits, a diode D used to connect in parallel with the resistor R. It is used for the purpose of bypass and thus giving improved dv/vt protection.

2. UNDERSTAND THE WORKING OF CONVERTERS, AC REGULATORS AND CHOPPERS.

EXPLAIN CONTROLLED RECTIFIERS TECHNIQUES (PHASE ANGLE, EXTINCTION ANGLE CONTROL & PWM)

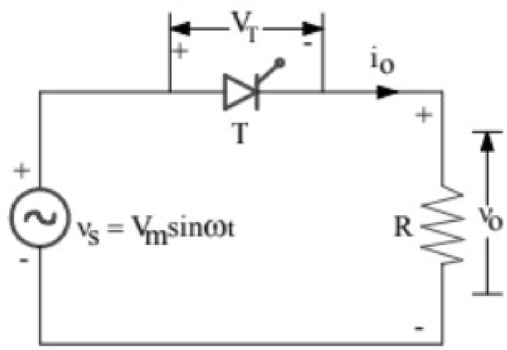
Definition:

Phase Control of SCR means having control on the phase relationship between the start of current through the SCR and source voltage.

Phase Control of SCR Explanation:

Whenever we talk of phase angle, we generally mean the angle of a sinusoidal quantity at any instant of time. Phase control of SCR means the phase angle (with reference to source voltage) where it is getting turned ON by the application of gate signal.

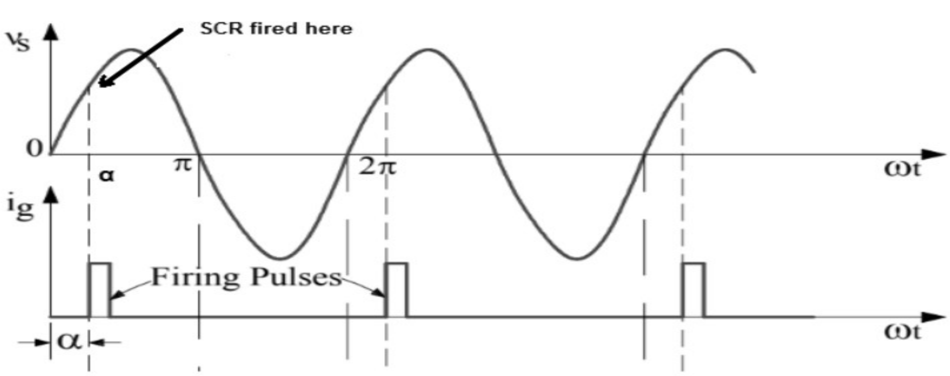
Let us understand the concept with the help of a simple circuit diagram as shown below.



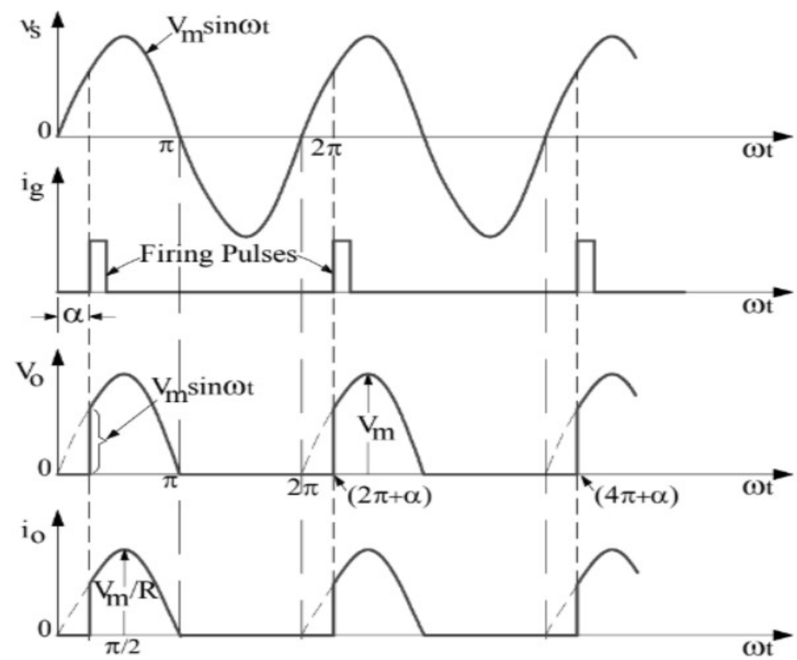
PHASE CONTROL OF SCR

In the above circuit, a thyristor (or SCR) T is connected to load R and voltage source vs. This SCR will not conduct until and unless it is forward biased and gate signal is applied. Application of gate signal is called firing.

During positive half cycle of supply voltage vs, the SCR is forward biased. If the thyristor T is fired (say at some phase angle α on the source voltage), it will become ON.



As SCR is now ON, it will start conducting. It will conduct from $\omega t = \alpha$ to π. Since load is resistive in nature, the load voltage v_0 and load current i_0 will follow the waveform of supply voltage. The load voltage, load current and supply voltage waveforms are shown in figure below. Compare the waveform of source voltage vs and load current i_0. You will observe that, the SCR is getting turned ON at a phase angle of α. Thus phase angle where thyristor T starts conducting is dependent on firing angle. If firing angle α is 0 degree, then load current and source voltage will be in phase whereas if α is 90 degree then load current will start when the source voltage is maximum. Thus the starting phase angle of load is controlled by firing angle.



At $\omega t = \pi$, thyristor T will get commutated as the load current becomes zero and SCR is reversed biased from $\omega t = \pi$ to 2π. This is known as Natural Commutation. We again need to fire SCR at $(2\pi+\alpha)$, $(4\pi+\alpha)$, $(6\pi+\alpha)$ and so on.

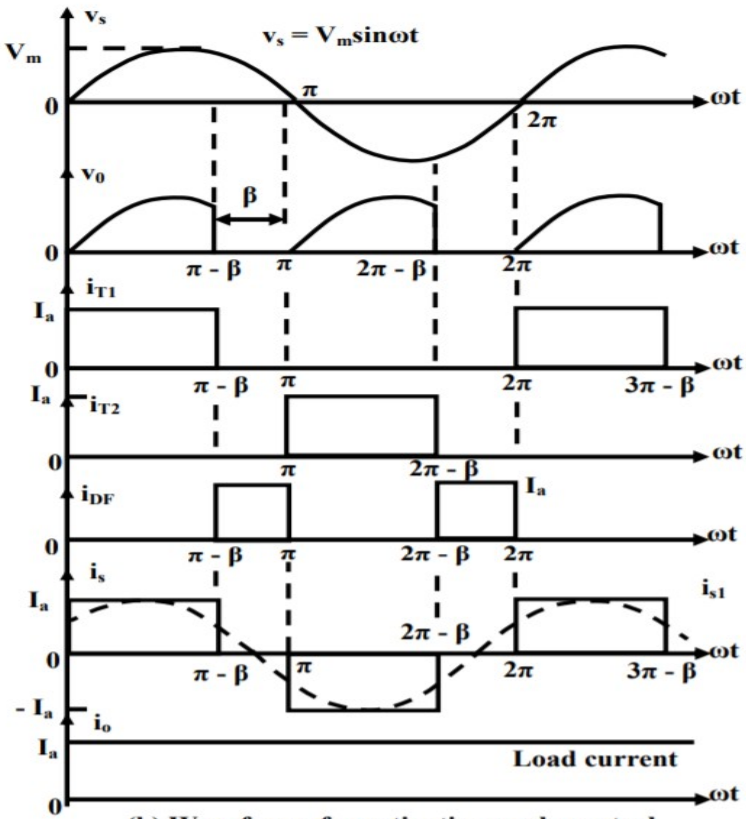
Extinction angle control

In power electronics extinction angle is one where thyristor gets switched off at desied angle. The output voltage is controlled by varying the extinction angle, β.

A gate turn-off thyristor (GTO) also may be used, in which case, it may be turned off by applying a short negative pulse to its gate, but is turned on by a short positive pulse, like a thyristor. In case of power transistor, the power transistor is turned on by applying a signal at the base, and turned off by withdrawing the signal at the base.

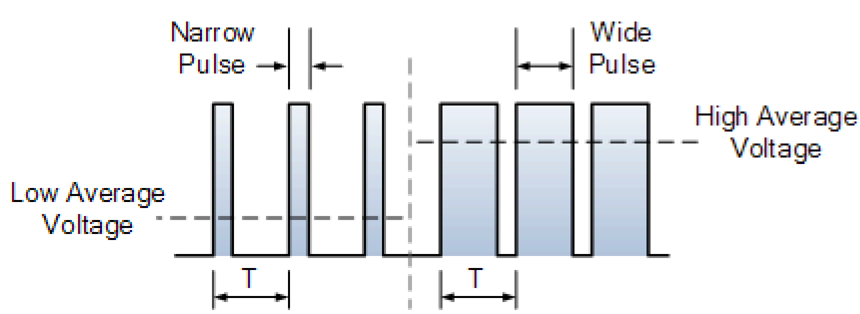
In extinction angle control, switch, S1 is turned on at $\omega t=0$, and then turned off by forced commutation at $\omega t = (\pi-\beta)$. The switch, S2 is turned on at $\omega t = \pi$, and then turned off at ($\omega t = 2\pi-\beta$). The output voltage is controlled by varying the extinction angle, β.

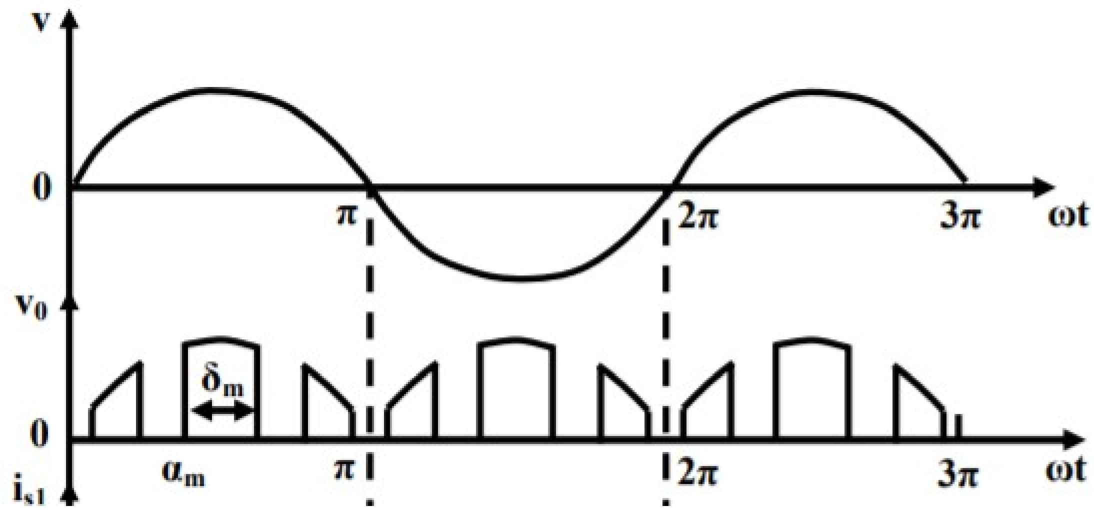
Fig. shows the waveforms for input voltage, output voltage, input current, and the current through thyristor switches



(b) Waveforms for extinction angle control

Pwm control





In Pulse Width Modulation (PWM) control, the converter switches are turned on and off several times during a half cycle, and the output voltage is controlled by varying the width of pulses. Fig. shows the input voltage, output voltage.

The average output voltage waveform and control pulses using PWM control is shown.

EXPLAIN SINGLE QUADRANT SEMI CONVERTER, TWO QUADRANT FULL CONVERTER AND DUAL CONVERTER

The phase controlled converters may be classified as semi-converter, full converter, dual converter. Depending on the input ac supply used they are classified as single phase and three phase converters.

Semi-converter: -A semi-converter is a one quadrant converter and it has one polarity of output voltage and current. It contains a mixture of diodes and thyristors allowing more limited control over the dc output voltage level than the full controlled rectifier. It is cheaper. It permits power flow from AC system to DC load. It is also known as half-wave controlled converter.

Full-converter: -A full-converter is a two-quadrant converter and the polarity of its output voltage can be either positive or negative. However, the output current of full-converter has one polarity only. Here power can be transmitted from AC side to DC side (conversion) and from DC side to AC side (inversion). It uses only thyristor as rectifying elements.

Dual-converter: -If two full converters are connected back to back they form a dual converter. It can operate in four quadrants and both the output voltage and current can be either positive or negative. Normally these are used in high power applications.

WORKING OF SINGLE-PHASE HALF WAVE CONTROLLED CONVERTER WITH RESISTIVE AND R-L LOADS.

In this article, we will see the analysis of Single Phase Half Wave Controlled Rectifier with Resistive (R) Load as shown in Figure 1. V_s is supply and 'i_s'is the source current. V_T and 'i_T' is the SCR voltage and current respectively. V_o and 'i_o'is the load voltage and current respectively.

$V_s = V_m \sin(\omega t)$

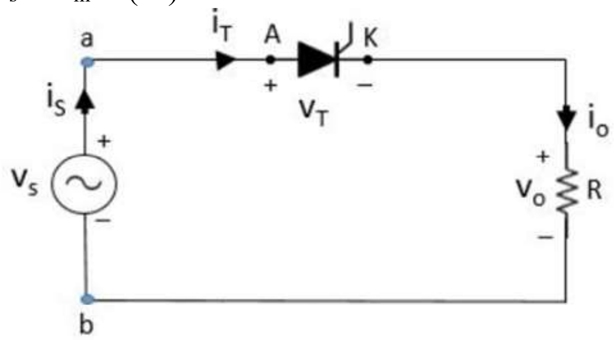


Fig. 1

Step-1: Write KVL in the given circuit

$v_s - v_T - v_o = 0$

$v_{ab} = v_s = v_T + v_o$

$v_{ab} = v_s = V_m \sin(\omega t)$

$v_{ba} = - v_{ab} = - V_m \sin(\omega t)$

Step-2: Device working status

In the positive half cycle, SCR is forward biased and is turned ON at $\omega t = \alpha$ by giving a firing pulse (triggering pulse). The angle **α** is called delay angle or firing angle. It is measured from the instant the SCR has become forward biased.

At $\omega t = \pi$, the current through the SCR falls to zero i.e. below zero and simultaneously a reverse voltage appears across SCR and it turns OFF. Since the line voltage is used for commutation, it is also known as line commutation converter.

1). $0 < \omega t < \alpha$

SCR is in forward blocking mode and SCR is OFF.

$i_o = i_s = i_T = 0$ Amp

$v_o = 0$ V

$v_T = v_{ab} = V_m \sin(\omega t)$

2). $\alpha < \omega t < \pi$

SCR is in forward conduction mode and SCR is ON.

$v_T = 0$ V

$i_o = i_s = i_T = (V_m \sin(\omega t))/R$

$v_o = v_{ab} = V_m \sin(\omega t)$

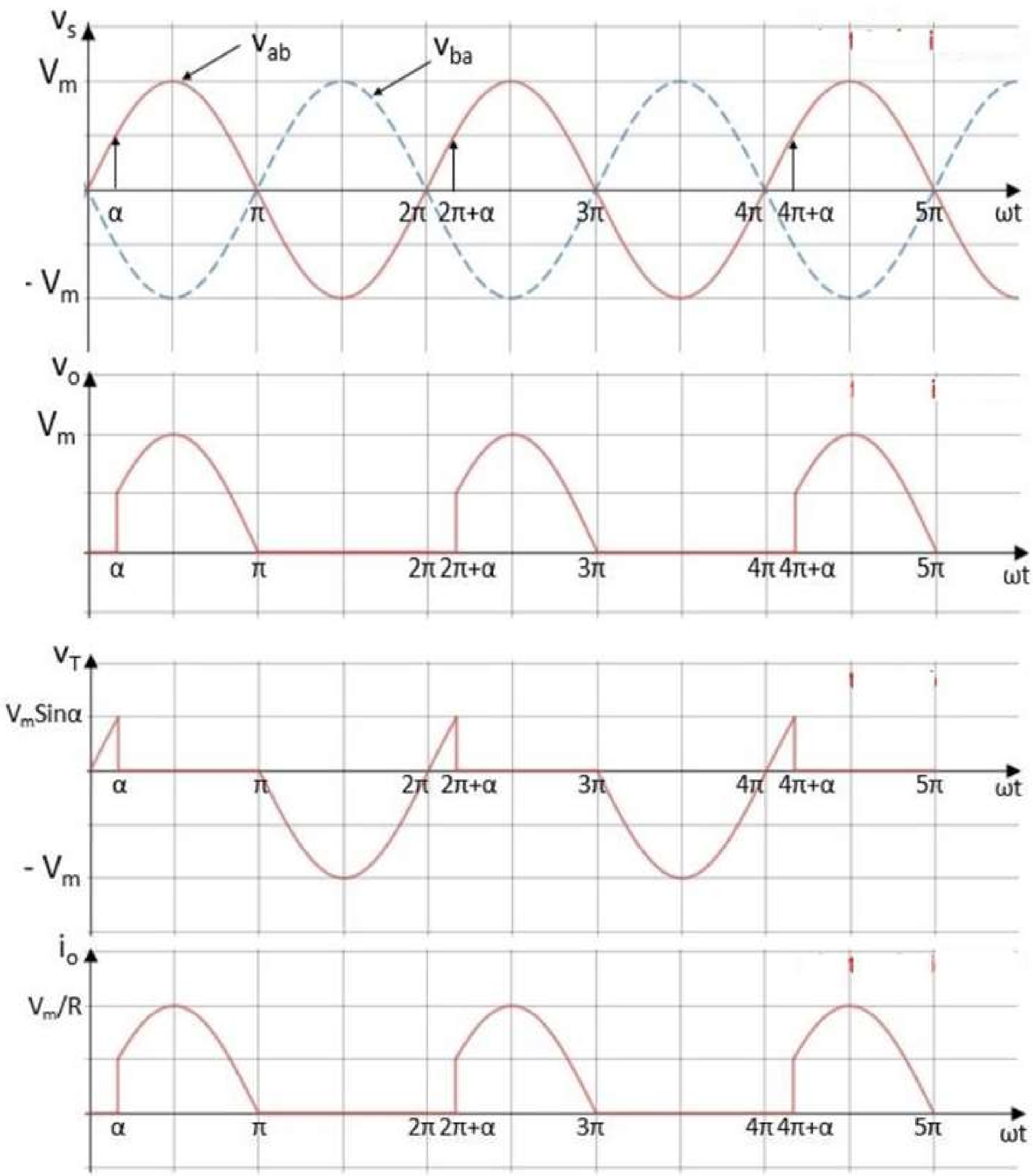


Fig. 2

3). $\pi < \omega t < 2\pi$

SCR is in forward blocking mode and SCR is OFF.

$i_o = i_s = i_T = 0$ Amp

$v_o = 0$ V

$v_T = v_{ab} = V_m \sin(\omega t)$

4). $2\pi + \alpha < \omega t < 3\pi$

SCR is in forward conduction mode and SCR is ON.

$v_T = 0$ V
$i_o = i_s = i_T = (V_m \sin(\omega t))/R$
$v_o = v_{ab} = V_m \sin(\omega t)$

The waveforms for load voltage (v_o), and current (i_o), SCR voltage (v_T) is shown in figure 2. The waveforms of supply current and SCR current is same as load current.
Note: For a resistive load, v_o and Io waveform will be same in nature except in magnitude.

Step-3:
1). Firing angle (α) = α

2). Extinction angle (β) = π

3). Conduction angle (γ) = $\beta - \alpha = \pi - \alpha$

4). Conduction time (t_c) = $\gamma/\omega = (\pi - \alpha)/\omega$
5). Circuit turn off time ($t_{ckt\text{-}off}$) = $(2\pi - \pi)/\omega = \pi/\omega$
6). Peak Inverse voltage (PIV) = V_m

Step-4:
Average Values
The average output voltage $V_{o(avg)}$ across load R is given by

$$V_{o(avg)} = \frac{1}{2\pi}\int_{\alpha}^{2\pi+\alpha} v_o(t)\ d\omega t$$
$$= \frac{1}{2\pi}\left[\int_{\alpha}^{\pi} V_m \sin(\omega t)\ d\omega t + \int_{\pi}^{2\pi+\alpha} 0\ d\omega t\right]$$
$$= \frac{1}{2\pi}\left[-V_m[Cos(\omega t)]_{\alpha}^{\pi}\right]$$
$$V_{o(avg)} = \frac{V_m}{2\pi}(1 + Cos\alpha)$$

The average output current $I_{o(avg)}$ through the load R is given by

$$I_{o(avg)} = \frac{V_{o(avg)}}{R}$$
$$I_{o(avg)} = \frac{V_m}{2\pi R}(1 + Cos\alpha)$$

RMS values

$$V_{o(rms)} = \sqrt{\frac{1}{2\pi}\int_{\alpha}^{2\pi+\alpha}[V_m \sin(\omega t)]^2\, d\omega t}$$

$$I_{o(rms)} = \frac{V_{o(rms)}}{R}$$

Single Phase Half Wave Controlled Rectifier with RL LOAD

The single phase half-wave controlled rectifier with inductive-load is shown in Fig.1.a The wave shapes for voltage and current in case of an inductive load are given in Fig.1.b. The load is assumed to be highly inductive.

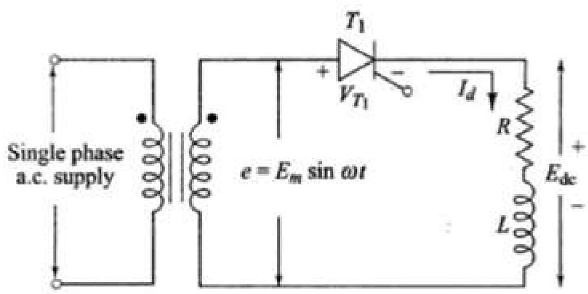


Fig.1(a) *Half-wave controlled rectifier with R-L load*

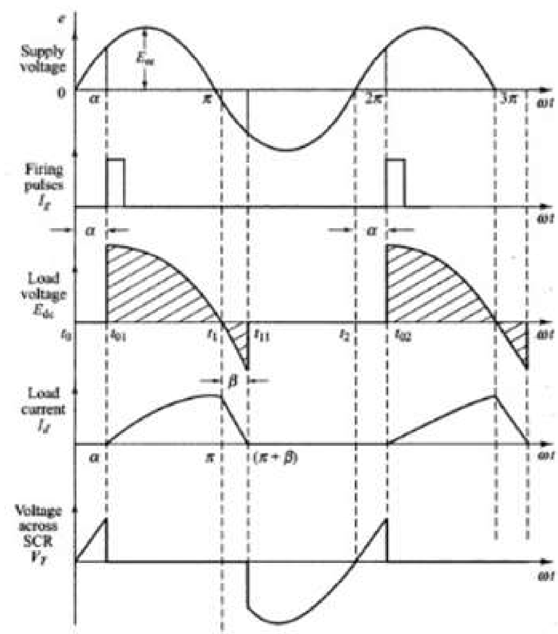


Fig.1(b) *Waveforms for a half-wave controlled rectifier with RL load*

The operation of the circuit on inductive loads changes slightly. Now at instant t01 , when the thyristor is triggered, the load-current will increase in a finite-time through the inductive load. The supply voltage from this instant appears across the load. Due to inductive load, the increase in current is gradual. Energy is stored in inductor during time t01 to t1. At t1, the supply voltage reverses, but the thyristor is kept conducting. This is due to the fact that current through the inductance cannot be reduced to zero.

During negative-voltage half-cycle, current continues to flow till the energy stored in the inductance is dissipated in the load-resistor and a part of the energy is fed-back to the source. Hence, due to energy stored in inductor, current , current continuous to flow up to instant t1 at instant, t11, the load-current is zero and due to negative supply voltage, thyristor turns-off.

At instant t02, when again pulse is applied, the above cycle repeats. Hence the effect of the inductive load is increased in the conduction period of the SCR.

The half-wave circuit is not normally used since it produces a large output voltage ripple and is incapable of providing continuous load-current.

The average value of the load-voltage can be derived as:

Here, it has been assumed that in negative half-cycles, the SCR conducts for a period of α

$$E_{dc} = \frac{E_m}{2\pi}[-\cos\omega t]_{\alpha}^{\pi+\alpha}$$

$$E_{dc} = \frac{E_m}{\pi}\cos\alpha$$

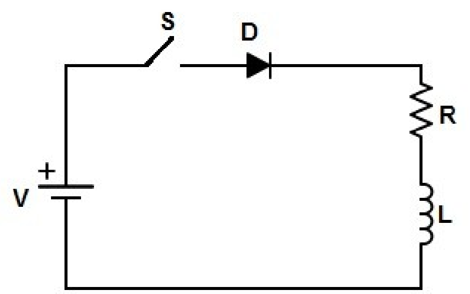
From the above equation, it is clear that the average load-voltage is in case of inductive load.

This is due to the conduction of SCR in negative cycle.

UNDERSTAND NEED OF FREEWHEELING DIODE.

A freewheeling diode is basically a diode connected across the inductive load terminals to prevent the development of high voltage across the switch. When the inductive circuit is switched off, this diode gives a short circuit path for the flow of inductor decay current and hence dissipation of stored energy in the inductor. This diode is also called Flywheel or Flyback diode

Working Principle of Freewheeling Diode:



When switch S is closed, the steady state current I through the circuit is (V/R) and hence the stored energy in inductor is $(LI^2)/2$. When this switch S is opened, the current will suddenly decay to zero from steady value I = (V/R). Due to this sudden decay of current, a high reverse voltage (as per lenz's law) equal to L(di/dt) will appear across the inductor terminals and hence across the diode and switch. This will lead to sparking across the switch contacts. If this reverse voltage exceeds the Peak Inverse Voltage of diode, then it may get damage. To avoid such occurrences, a diode, called freewheeling or flyback diode is connected across the inductive load RL as shown in figure below.

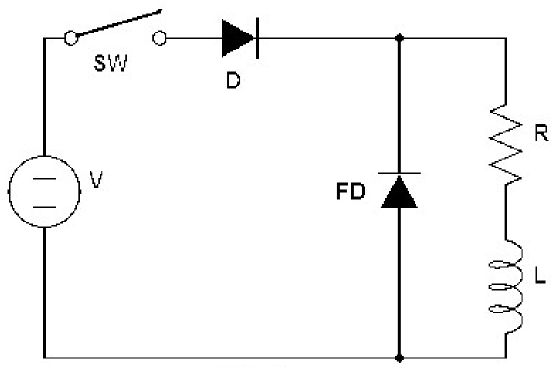
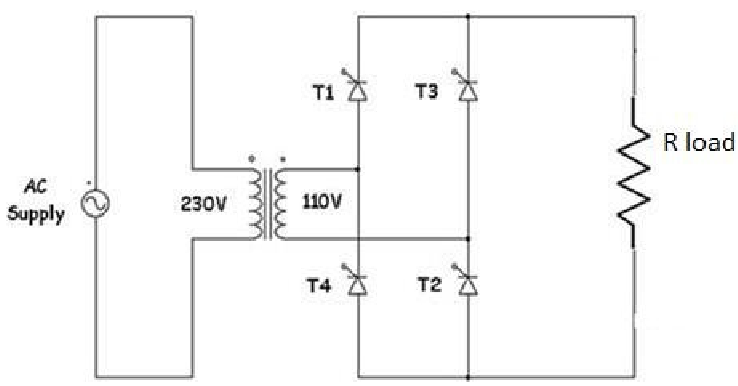


Fig - 1

Thus, the main circuit current is transferred to the circuit consisting of freewheeling diode FD, R and L as shown in above figure. In this new circuit, the current will exponentially decay to zero Thus we see that, freewheeling diode dissipates the stored energy in inductor by providing a short circuit path. It also provides a shorted path for exponentially decay of circuit current. Thus high voltage is not induced. Therefore, the switches and diode is protected from the high voltage.

WORKING OF SINGLE PHASE FULLY CONTROLLED CONVERTER WITH RESISTIVE AND R- L LOADS.

WORKING OF SINGLE PHASE FULLY CONTROLLED CONVERTER WITH RESISTIVE LOAD



Full wave bridge rectifier circuit with resistive load.

Figure above shows a full-wave controlled bridge rectifier circuit with a resistive load. In this circuit, diagonally opposite pairs of SCRs turn on and off together.
During the positive half-cycle of the input .voltage, SCRl, SCR2 is forward-biased. If we apply the gate signal at a SCRl and SCR2 turns on. The output Voltage (Vo) follows the input voltage. The load current (io = vo/R|) has the same waveform as the load voltage. At π, when the current through SCR1and SCR4 becomes zero, it turns off naturally.

During the negative half-cycle; SCR3 and SCR4 is forward biased. SCR3 and SCR4 is fired at $(\pi + \alpha)$. The output voltage again follows the input voltage."The current through SCR4 and SCR3 becomes zero at 2 π; and it turns off.

SCRl, SCR2 is fired again at (2π+α), SCR3 and SCR4 at (3π+α), and the cycle repeats.

Figure shows the resulting voltage and current waveforms.. The average DC output voltage can be controlled from zero to its maximum value by varying the firing angle.

The SCRs are controlled and fire in pairs with a delay angle of α. The current and voltage waveform become full-wave, as shown in figure below

(a) Average output voltage (V_{dc})

$$V_{dc} = \frac{1}{\pi}\int_{\alpha}^{\pi} V_m \sin\omega t . d\omega t$$

$$V_{dc} = \frac{V_m}{\pi}(1+\cos\alpha)$$

(b) RMS output voltage (V_{0rms})

$$V_{0rms} = \left[\frac{1}{\pi}\int_{\alpha}^{\pi} V_m^2 \sin^2 \omega t . d\omega t\right]^{1/2} = V_m\left[\frac{\pi-\alpha}{2\pi} + \frac{\sin 2\alpha}{4\pi}\right]^{1/2}$$

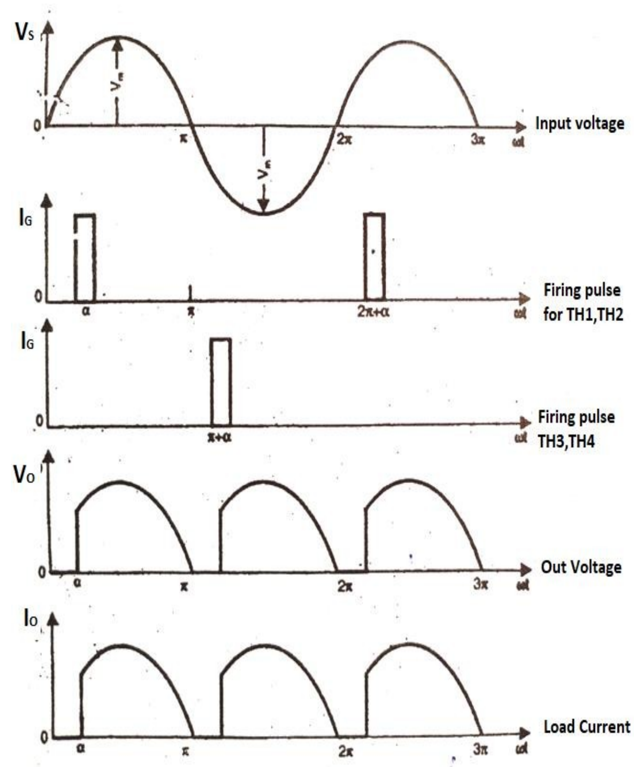
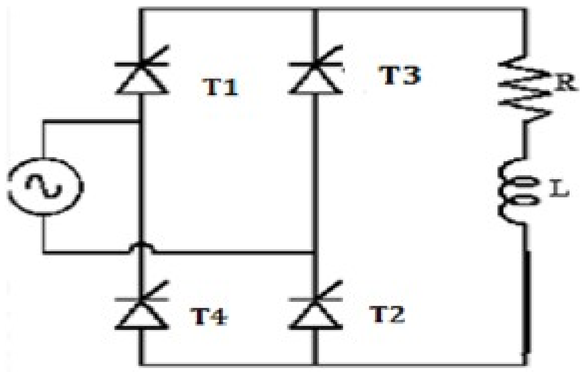


Fig. Voltage and current waveforms

SINGLE PHASE FULL WAVE CONVERTER WITH RL LOAD



Operation of this mode can be divided between four modes

Mode 1 (α to π)

• In positive half cycle of applied ac signal, SCR's T1 & T2 are forward bias & can be turned on at an angle α.

• Load voltage is equal to positive instantaneous ac supply voltage. The load current is positive, ripple free, constant and equal to Io.

Due to positive polarity of load voltage & load current, load inductance will store energy.

Mode 2 (π to π+α)

• At wt=π, input supply is equal to zero & after π it becomes negative. But inductance opposes any change through it.

• In order to maintain a constant load current & also in same direction. A self induced emf appears across 'L' as shown.

• Due to this induced voltage, SCR's T1 & T2 are forward bias in spite the negative supply voltage.

The load voltage is negative & equal to instantaneous ac supply voltage whereas load current is positive.

• Thus, load acts as source & stored energy in inductance is returned back to the ac supply.

Mode 3 (π+α to 2π)

• At wt=π+α SCR's T3 & T4 are turned on & T1, T2 are reversed bias.

• Thus, process of conduction is transferred from T1, T2 to T3, T4.

• Load voltage again becomes positive & energy is stored in inductor

• T3, T4 conduct in negative half cycle from (π+α) to 2π with positive load voltage & load current energy gets stored

Mode 4 (2π to 2π+α)

• At wt=2π, input voltage passes through zero.

• Inductive load will try to oppose any change in current if in order to maintain load current constant & in the same direction.

• Induced emf is positive & maintains conducting SCR's T3 & T4 with reverse polarity also.

• Thus VL is negative & equal to instantaneous ac supply voltage. Whereas load current continues to be positive.

• Thus load acts as source & stored energy in inductance is returned back to ac supply

• At wt=α or 2π+α, T3 & T4 are commutated and T1, T2 are turned on.

(a) Average output voltage (V_{dc})

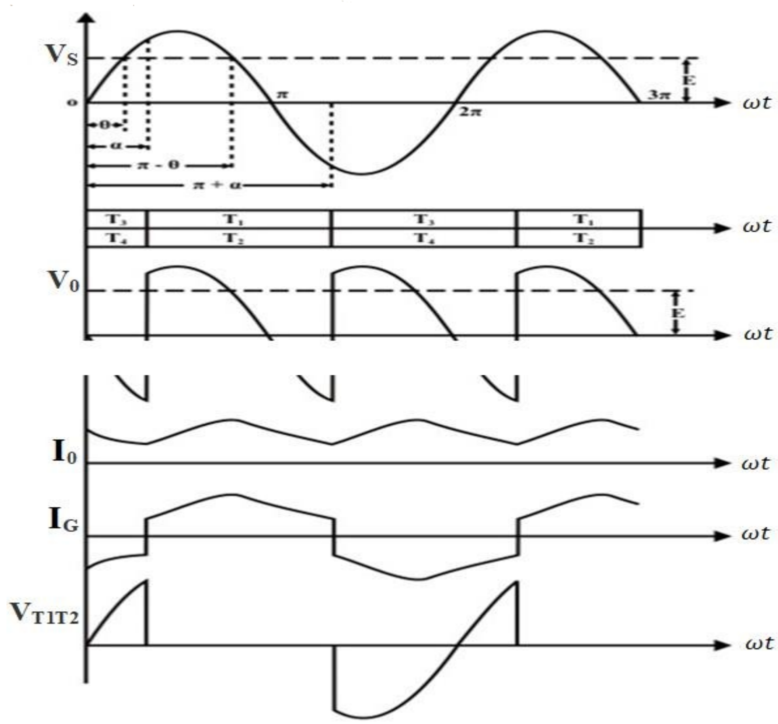
$$V_{dc} = \frac{1}{\pi}\int_{\alpha}^{\pi+\alpha} V_m \sin\omega t.d(\omega t) = \frac{V_m}{\pi}\left|-\cos\omega t\right|_{\alpha}^{\pi+\alpha}$$

$$= \frac{V_m}{\pi}\left|\cos\omega t\right|_{\pi+\alpha}^{\alpha} = \frac{V_m}{\pi}\left[\cos\alpha - (-\cos\alpha)\right]$$

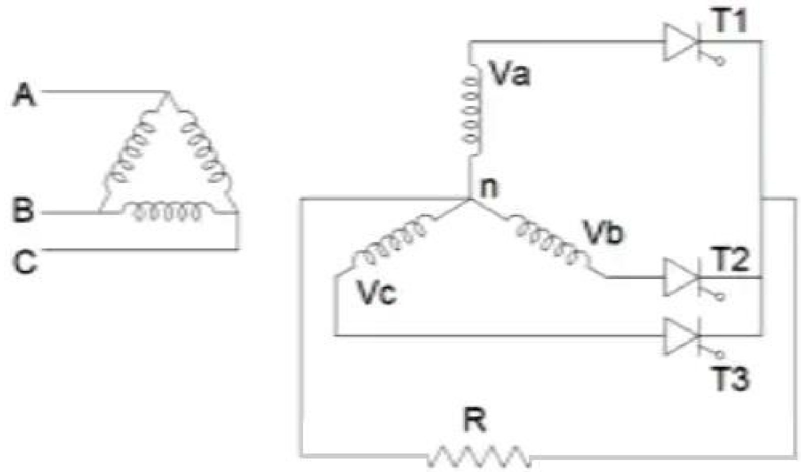
$$\boxed{V_{dc} = \frac{2V_m}{\pi}\cos\alpha}$$

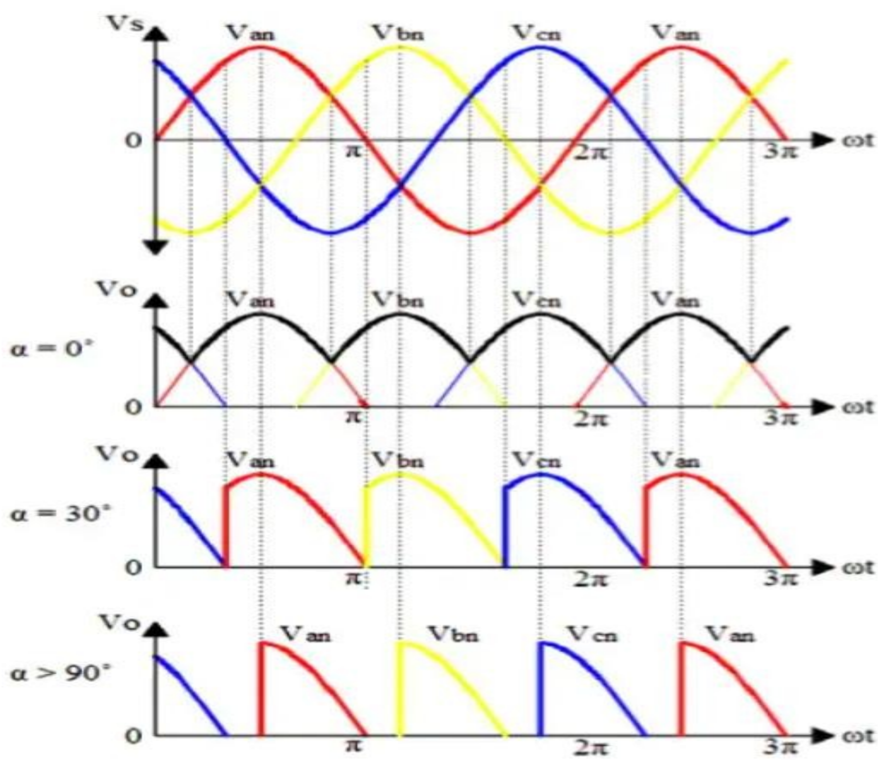
(b) RMS output voltage (V_{orms})

$$V_{orms} = \left[\frac{1}{\pi}\int_{\alpha}^{\pi+\alpha} V_m^2 \sin^2 \omega t d(\omega t)\right]^{1/2} = V_m\left[\frac{1}{2} + \frac{\sin 2\alpha}{4\pi}\right]^{1/2}$$



WORKING OF THREE-PHASE HALF WAVE CONTROLLED CONVERTER WITHRESISTIVE LOAD



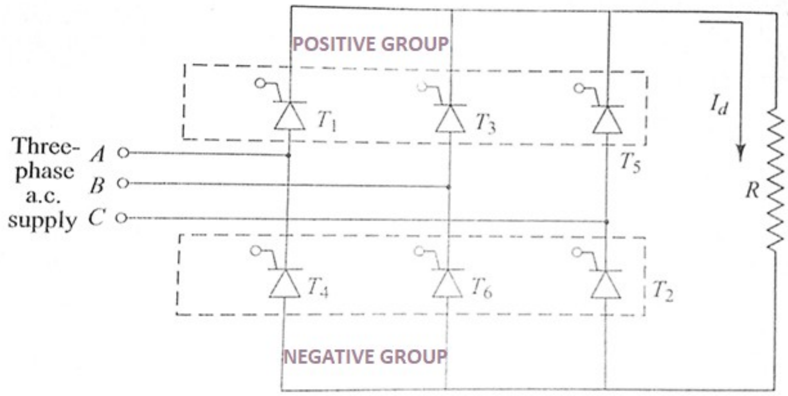


WORKING PRINCIPLE

- The circuit consists of a delta star transformer and 3 thyristors T1, T2, T3 which are connected on the secondary star connected winding and a resistive load.
- When Va is positive, T1 becomes forward biased and conducts when triggered. During the negative cycle of Va, T1 turns OFF.

- When VB is positive, T2 becomes forward biased and conducts when triggered. During the negative cycle of Vb, T2 turns OFF.
- Similarly T3 conducts only during the positive cycles of Vc when triggered and turns OFF during negative cycle of Vc
- The above figure shows waveforms for different triggering angles.
- The average output voltage can be varied by varying the firing angles of the thyristors.
- The waveform shows the output voltage for various firing angles. In the waveform, Va is denoted as Van, Vb as Vbn, Vc as Vcn.

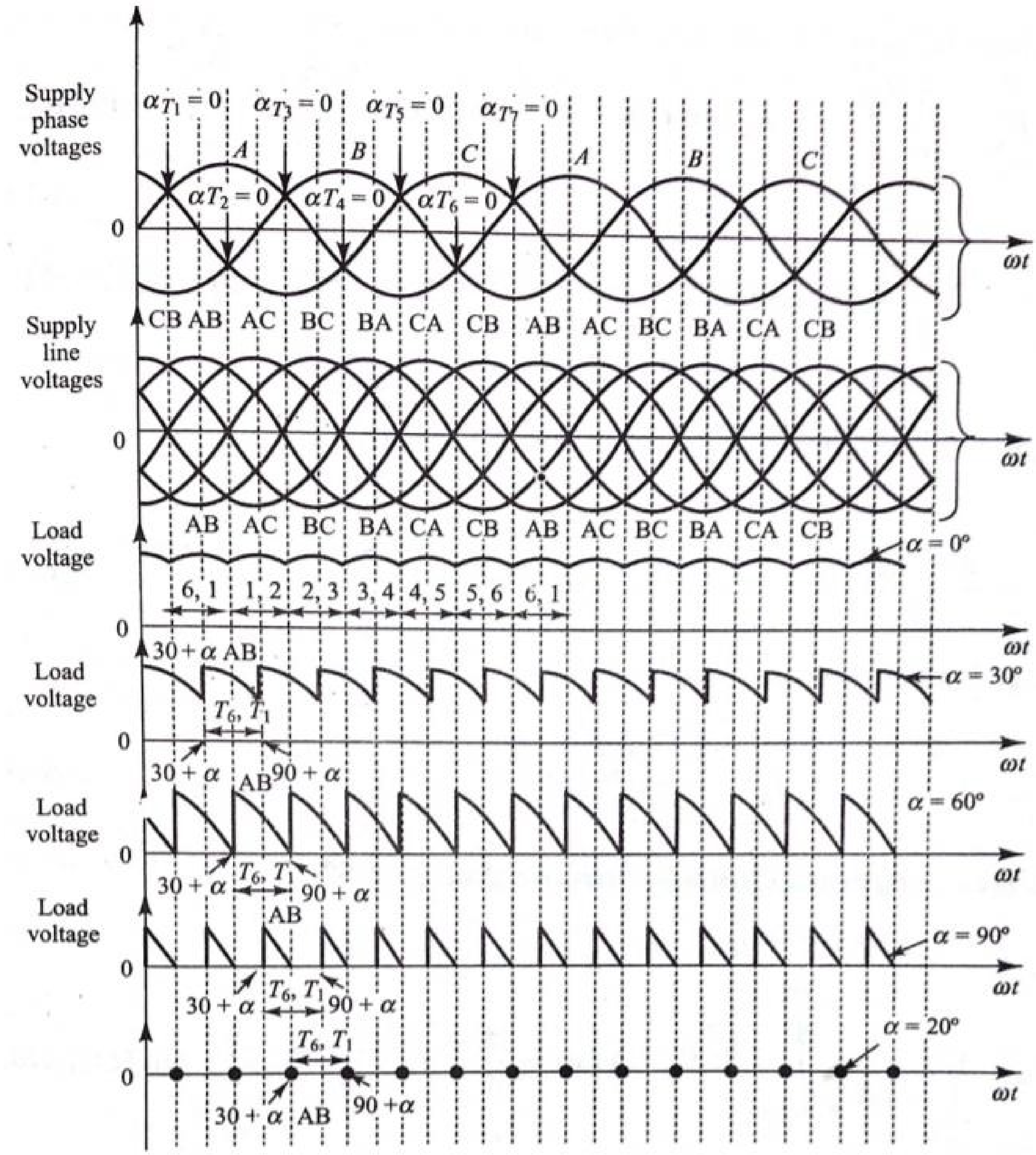
WORKING OF THREE PHASE FULLY CONTROLLED CONVERTER WITHRESISTIVE LOAD.



3 φ FULL CONVERTER

Three phase full wave controlled rectifier with resistive load is shown above
The circuit diagram consists of 6 SCRS, SCR T1, T3, T5 forms positive group and SCR T4 T6 T2 forms negative group.
The positive groups of SCRs are turned on when the supply voltage is positive and the negative groups of SCRs are turned on when the supply voltage is negative.
Operation
Each device is triggered at a desired firing angle Alpha (α)
Each SCR conducts for 120 degree
SCRs are triggered in the sequence T1 T2 T3 T4 T5 T6
Each SCRs conducts in two pair and each pair conducts for 60 degree
When two SCRs are conducting one from positive group and one from negative group the corresponding line voltage is applied across the load
The six pairs of thyristors which conducts are (T6 T1) (T1 T2) (T2 T3) (T3 T4) (T5 T6)
FIRING SEQUENCE OF SCR

S.No.	ωt	Incoming SCR	Conducting pair	Outgoing SCR	Line voltage across the load
1.	$30° + \alpha$	T_1	(T_6, T_1)	T_5	E_{AB}
2.	$90° + \alpha$	T_2	(T_1, T_2)	T_6	E_{AC}
3.	$150° + \alpha$	T_3	(T_2, T_3)	T_1	E_{BC}
4.	$210° + \alpha$	T_4	(T_3, T_4)	T_2	E_{BA}
5.	$270° + \alpha$	T_5	(T_4, T_5)	T_3	E_{CA}
6.	$330° + \alpha$	T_6	(T_5, T_6)	T_4	E_{CB}



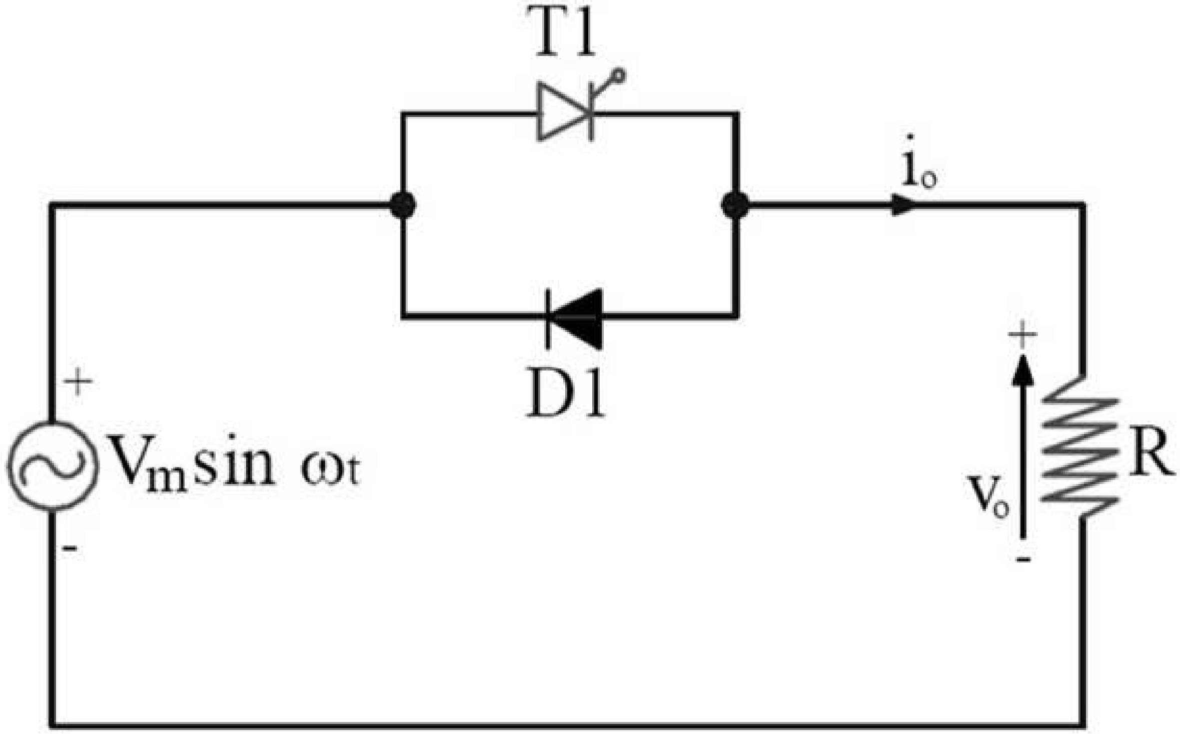
VOLTAGE WAVEFORM AND CONDUCTION OF THYRISTORS FOR 3-PHASE FULL CONVERTER

WORKING OF SINGLE PHASE AC REGULATOR

Single Phase AC Voltage regulator is a device which converts fixed single phase alternating voltage directly to a variable alternating voltage without a change in frequency. The input and output of the device is single phase. There are two types of single phase AC Voltage Controller i.e. **Single phase Half Wave and Single phase Full Wave regulator.**

Single Phase Half Wave AC Voltage Controller:

A single phase half wave AC voltage controller comprises of a thyristor connected in anti-parallel with a power diode. The circuit diagram is shown in figure below.



The load is assumed resistive for the sake of simplicity. The input source is $V_m Sin\omega t$.

For the positive half cycle of input source, thyristor T1 is forward biased and hence it is able to conduct provided gate signal is applied. This means that T1 will remain OFF until gate signal is applied.

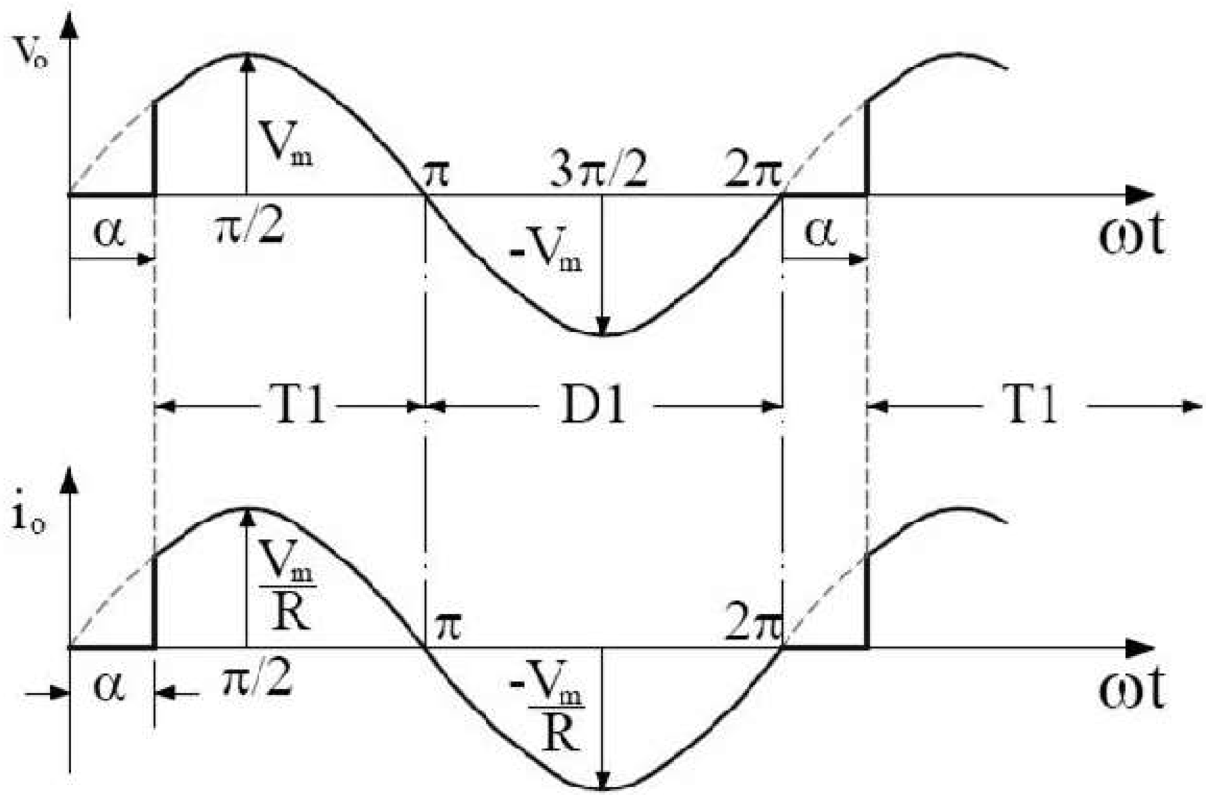
Now suppose, at some angle α (called the firing angle), thyristor T1 is gated. As soon as T1 is fired / gated, it starts conducting and hence, load gets directly connected to the source. This makes load voltage $V_o = V_m Sin\alpha$ and load current $I_o = (V_m Sin\alpha / R)$ at the instant T1 is fired.

From $\omega t = \alpha$ to π, the load voltage and current follows the input voltage waveform $V_m Sin\omega t$ and $(V_m Sin\omega t / R)$ respectively.

After $\omega t = \pi$, thyristor T1 becomes reversed biased and the load current becomes zero (note that load voltage and current are in phase, hence as soon as load voltage becomes zero, load current also becomes zero) and hence thyristor T1 is commutated naturally.

After $\omega t = \pi$, diode D1 becomes forward biased and hence starts conducting. This makes load voltage & current to follow the supply voltage $V_m Sin\omega t$ and $(V_m Sin\omega t / R)$ respectively for the negative half cycle.

The output waveform for load voltage & current is shown below.



Following points may be noted from the above waveforms:

- By having a control on the firing angle α, the load voltage may be controlled. It may be seen from the output waveform that, there is no control on the negative half cycle of the input voltage. This is the reason, a single phase half wave AC voltage controller is also known as single phase unidirectional voltage controller.
- The positive and negative half cycle of the load voltage & current are not identical. As a result, DC component is introduced in the supply and load circuit which is undesirable.

Let us now calculate the rms value of load voltage and current. This will give us an idea of the magnitude of output voltage and current.

$RMS\ Load\ Voltage\ Vo =$

$$\sqrt{\frac{1}{2\pi}\int_{\alpha}^{2\pi}\{V_m\ Sin\omega t\}^2\, d(\omega t)}$$

$$= \frac{V_m}{2}\sqrt{\frac{1}{\pi}\{(2\pi - \alpha) + \frac{Sin2\alpha}{2}\}}$$

RMS Load Current = Vo / R

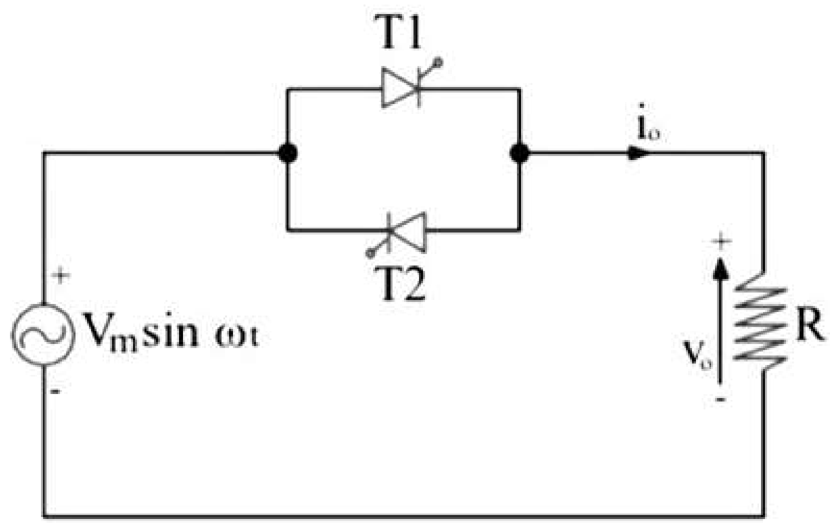
$Average\ Load\ Voltage =$

$$\frac{1}{2\pi}\int_{\alpha}^{2\pi} V_m\ Sin\omega t\, d(\omega t)$$

$$= \frac{V_m}{2\pi}(Cos\alpha - 1)$$

Single Phase Full Wave AC Voltage Controller:

A single phase full wave AC voltage controller comprises of two thyristor connected in anti-parallel. The circuit diagram is shown in figure below.



The load is assumed resistive for the sake of simplicity. The input source is V_mSinωt.

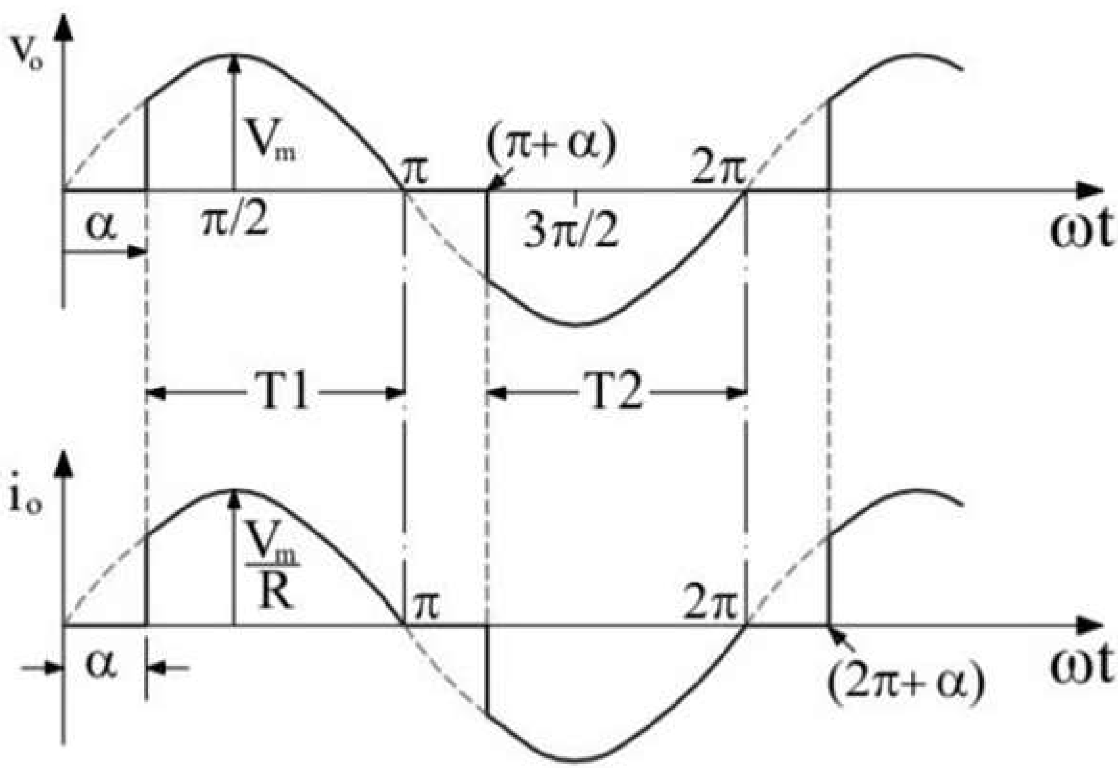
For the positive half cycle of input source, thyristor T1 is forward biased and hence it is able to conduct provided gate signal is applied. This means that T1 will remain OFF until gate signal is applied.
Now suppose, at some angle α (called the firing angle), thyristor T1 is gated. As soon as T1 is fired / gated, it starts conducting and hence, load gets directly connected to the source. This makes load voltage $V_o = V_m$Sinα and load current I_o = (V_mSinα / R) at the instant T1 is fired.

From wt = α to π, the load voltage and current follows the input voltage waveform V_mSinωt and (V_mSinωt / R) respectively.

At wt = π, the load voltage becomes zero and current, also, becomes zero. Since, thyristor T1 is reversed biased after ωt = π and current through it is zero, it gets naturally commutated.

At ωt = (π+α), forward biased thyristor T2 is gated. Hence, it conducts and connected load to the source. The load voltage now follows the negative envelop of the AC input supply and the load current does the same. Thus, the root mean square voltage may be controlled by having a control of firing angle. In this way, voltage control is achieved in AC voltage Controller.

The output waveform for load voltage & current is shown below.



It may be noted from the above waveform that the positive and negative half cycle of the load voltage & current are identical. As a result, DC component is not introduced in the supply and load circuit. This is the main advantage of single phase full wave AC voltage controller.
Single phase full wave AC voltage controller is also known as single phase bidirectional voltage controller. Let us now calculate the rms value of load voltage and current.

$RMS\ Load\ Voltage\ Vo =$

$$\sqrt{\frac{1}{\pi}\int_{\alpha}^{\pi}\{V_m\ Sin\omega t\}^2\ d(\omega t)}$$

$$= V_m\sqrt{\frac{1}{2\pi}\{(\pi-\alpha)\ +\frac{Sin2\alpha}{2}\}}$$

RMS Load Current = Vo / R

$Average\ Load\ Voltage = 0$

Single phase full wave voltage controllers are more suitable to practical circuits. It also overcomes the problem of dc component which is present in supply and load circuit of half wave voltage controller.

WORKING PRINCIPLE OF STEP UP & STEP DOWN CHOPPER.

A chopper uses high speed to connect and disconnect from a source load. A fixed DC voltage is applied intermittently to the source load by continuously triggering the power switch ON/OFF. The period of time for which the power switch stays ON or OFF is referred to as the chopper's ON and OFF state times, respectively.

Choppers are mostly applied in electric cars, conversion of wind and solar energy, and DC motor regulators.

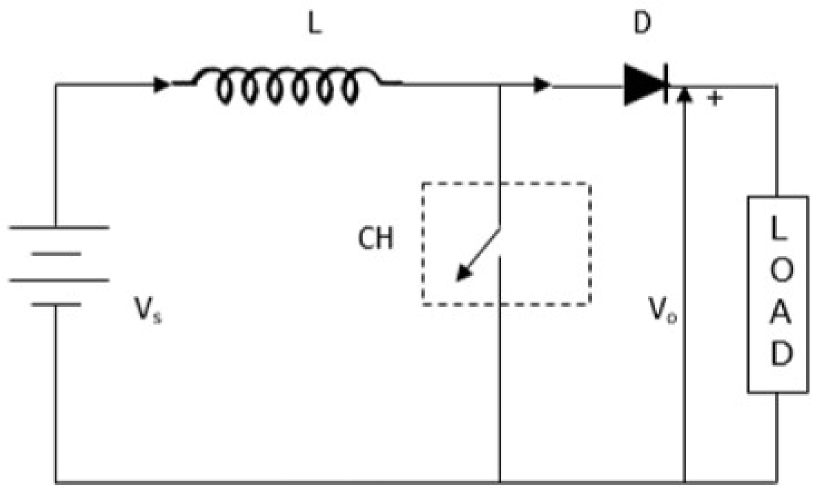
Classification of Choppers

Depending on the voltage output, choppers are classified as −

- Step Up chopper boost converter
- Step Down Chopper Buck converter
- Step Up/Down Chopper Buck−boost converter

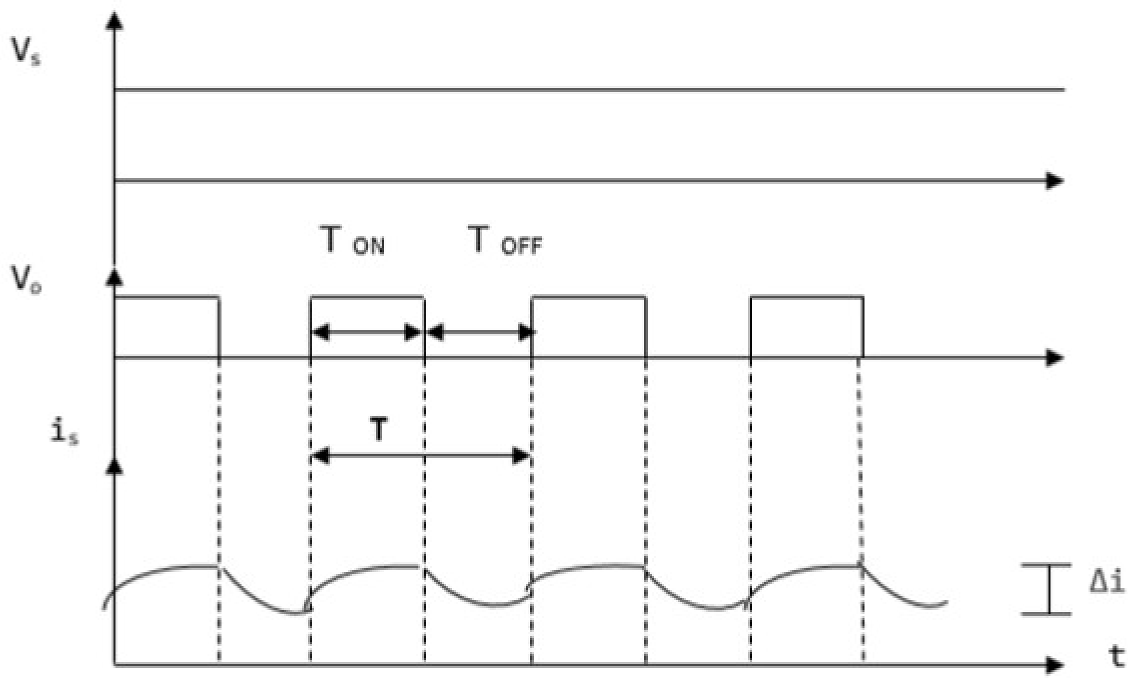
STEP UP CHOPPER.

The average voltage output (V_o) in a step up chopper is greater than the voltage input (V_s). The figure below shows a configuration of a step up chopper.



Current and Voltage Waveforms

V_0 average voltage output is positive when chopper is switched ON and negative when the chopper is OFF as shown in the waveform below.



Where

T_{ON} – time interval when chopper is ON

T_{OFF} – time interval when chopper is OFF

V_L – Load voltage

V_s – Source voltage

T – Chopping time period = $T_{ON} + T_{OFF}$

V_o is given by – $V_0 = \frac{1}{T}\int_0^{T_{ON}} V_S dt$

When the chopper CHCH is switched ON, the load is short circuited and, therefore, the voltage output for the period T_{ON} is zero. In addition, the inductor is charged during this time. This gives $V_S = V_L$

$$L\frac{di}{dt} = V_S, \qquad \frac{\Delta i}{T_{ON}} = \frac{V_S}{L}$$

Hence, $\Delta i = \frac{V_S}{L} T_{ON}$

Δi = is the inductor peak to peak current. When the chopper CHCH is OFF, discharge occurs through the inductor L. Therefore, the summation of the V_s and V_L is given as follows −

V0=VS+VL, VL=V0−VS

But $$L\frac{di}{dt} = V_0 - V_S$$

Thus, $$L\frac{\Delta i}{T_{OFF}} = V_0 - V_S$$

This gives, $$\Delta i = \frac{V_0 - V_S}{L} T_{OFF}$$

Equating Δi from ON state to Δi from OFF state gives –

$$\frac{V_S}{L} T_{ON} = \frac{V_0 - V_S}{L} T_{OFF} \ , \qquad V_S\left(T_{ON} + T_{OFF}\right) = V_0 T_{OFF}$$

$$V_0 = \frac{T V_S}{T_{OFF}} = \frac{V_S}{\frac{(T + T_{ON})}{T}}$$

This give the average voltage output as,

$$V_0 = \frac{V_S}{1 - D}$$

The above equation shows that V_o can be varied from V_S to infinity. It proves that the output voltage will always be more than the voltage input and hence, it boosts up or increases the voltage level.

STEP DOWN CHOPPER

This is also known as a buck converter. In this chopper, the average voltage output V_O is less than the input voltage V_S. When the chopper is ON, $V_O = V_S$ and when the chopper is off, $V_O = 0$

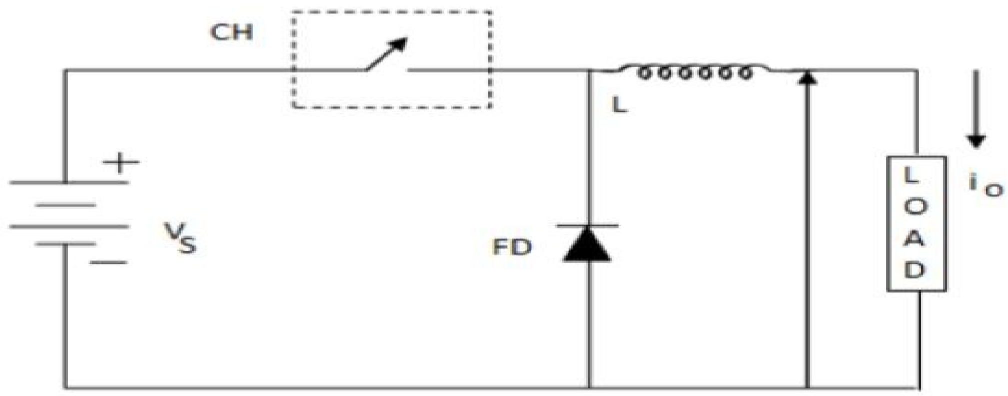
When the chopper is ON –

$$V_S = (V_L + V_0), \quad V_L = V_S - V_0, \quad L\frac{di}{dt} = V_S - V_0, \quad L\frac{\Delta i}{T_{ON}} = V_s + V_0$$

Thus, peak-to-peak current load is given by,

$$\Delta i = \frac{V_s - V_0}{L} T_{ON}$$

Circuit Diagram



Where **FD** is free-wheel diode.

When the chopper is OFF, polarity reversal and discharging occurs at the inductor. The current passes through the free-wheel diode and the inductor to the load. This gives,

$$L\frac{di}{dt} = V_0 \ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots (i)$$

Rewritten as – $L\frac{\Delta i}{T_{OFF}} = V_0$

$$\Delta i = V_0 \frac{T_{OFF}}{L} \ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots (ii)$$

Equating equations i and ii gives;

$$\frac{V_S - V_0}{L} T_{ON} = \frac{V_0}{L} T_{OFF}$$

$$\frac{V_S - V_0}{V_0} = \frac{T_{OFF}}{T_{ON}}$$

$$\frac{V_S}{V_0} = \frac{T_{ON} - T_{OFF}}{T_{ON}}$$

The above equation gives;

$$V_0 = \frac{T_{ON}}{T} V_S = DV_S$$

Equation i give –

$$\Delta i = \frac{V_S - DV_S}{L} DT \text{ , from } D = \frac{T_{ON}}{T}$$

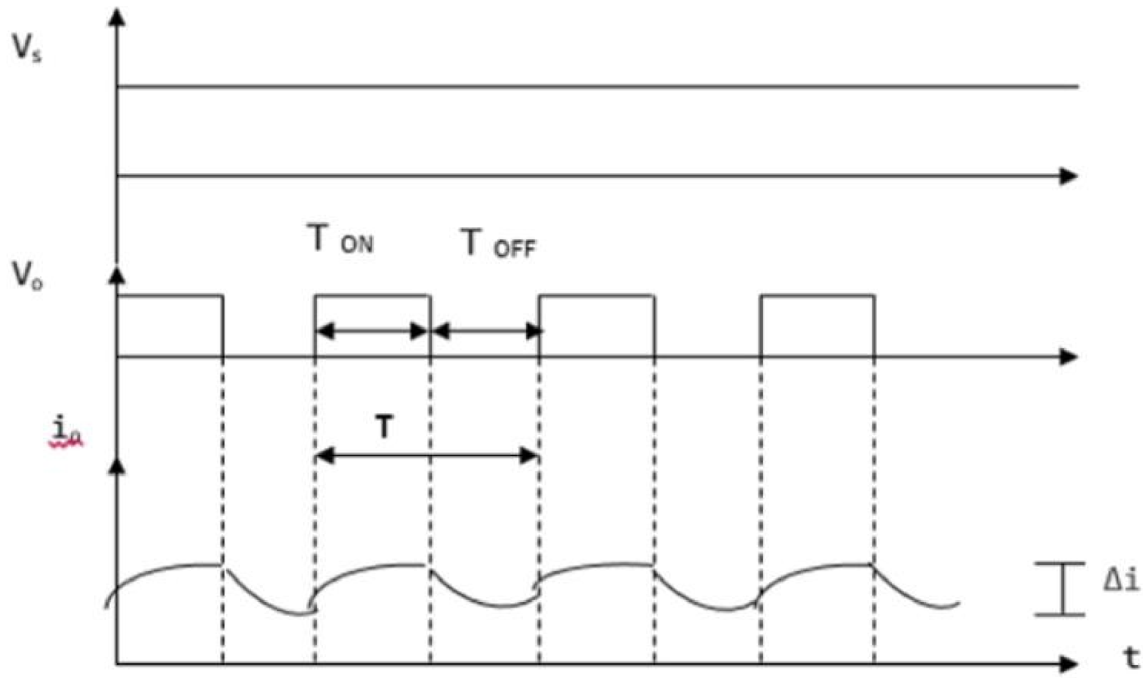
$$= \frac{V_S - (1-D)D}{Lf}$$

$$f = \frac{1}{T} = \text{ chopping frequency}$$

Current and Voltage Waveforms

The current and voltage waveforms are given below −

For a step down chopper the voltage output is always less than the voltage input. This is shown by the waveform below.



CONTROL MODES OF CHOPPER

In DC-DC converters, the average output voltage is controlled by varying the alpha (α) value.
This is achieved by varying the Duty Cycle of the switching pulses.
Duty cycle can be varied usually
in 2 ways:

1. Time Ratio Control
2. Current Limit Control

In this post we shall look upon both the ways of varying the duty cycle.
Duty Cycle is the ratio of 'On Time' to 'Time Period of a pulse'.

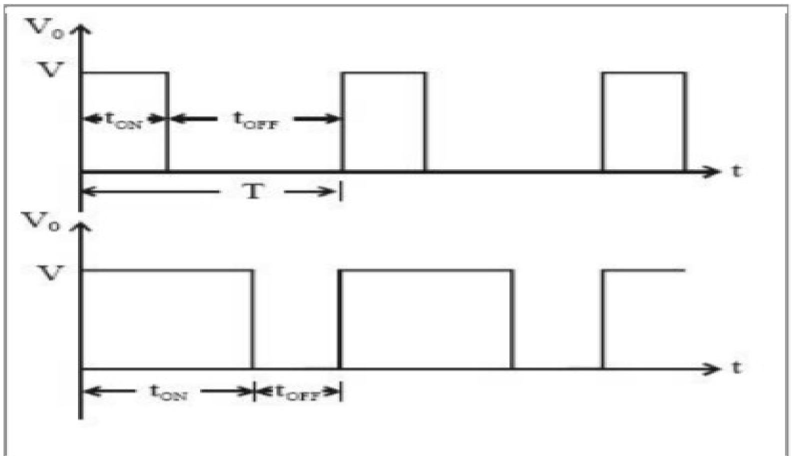
Time Ratio Control:

As the name suggest, here the time ratio (i.e. the duty cycle ratio Ton/T) is varied.
This kind of control can be achieved using 2 ways:

- Pulse Width Modulation (PWM)
- Frequency Modulation Control (FMC)

Pulse Width Modulation (PWM):

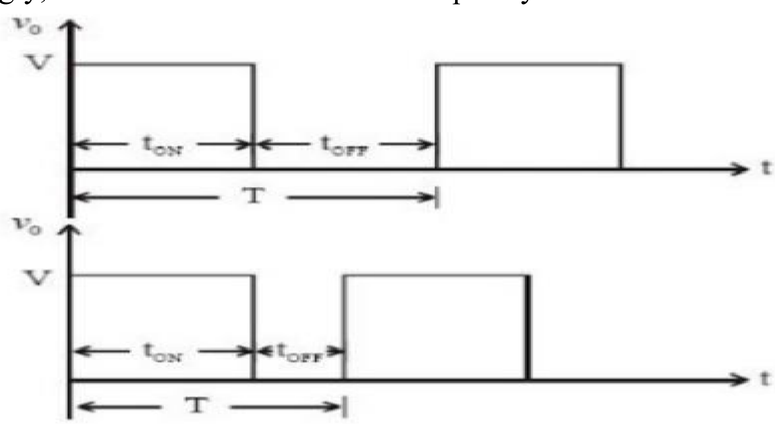
In this technique, the time period is kept constant, but the 'On Time' or the 'OFF Time' is varied. Using this, the duty cycle ratio can be varied. Since the ON time or the 'pulse width' is getting changed in this method, so it is popularly known as Pulse width modulation.



Pulse Width Modulation

Frequency Modulation Control (FMC)

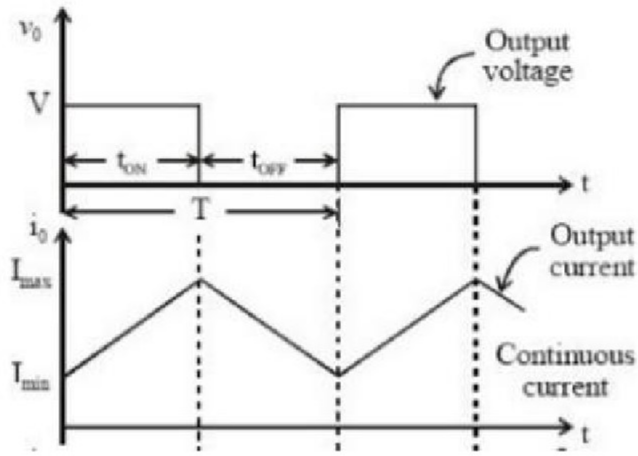
In this control method, the ‘Time Period’ is varied while keeping either of ‘On Time’ or ‘OFF time’ as constant. In this method, since the time period gets changed, so the frequency also changes accordingly, so this method is known as frequency modulation control.



Frequency Modulation Control

Current Limit Control:

As is obvious from its name, in this control strategy, a specific limit is applied on the current variation. In this method, current is allowed to fluctuate or change only between 2 values i.e. maximum current (I max) and minimum current (I min). When the current is at minimum value, the chopper is switched ON. After this instance, the current starts increasing, and when it reaches up to maximum value, the chopper is switched off allowing the current to fall back to minimum value. This cycle continues again and again.



Current Limit Control

OPERATION OF CHOPPER IN ALL FOUR QUADRANTS

CHOPPERCONFIGURATION AND QUADRANT OF OPERATION

The configuration of chopper is done according to direction of output voltage and output current.

Type A Chopper

The output voltage and current both are positive in the Type A chopper.

It works in the first quadrant.

The Type A chopper works as forward drive.

Type B Chopper

The output voltage is positive but output current negative in the Type B chopper.

The power flow from load to supply side. It works in second quadrant.

This Type of chopper is used in the regenerative braking of DC Motor.

Type C Chopper

The load current is either positive or negative but load voltage remains positive in the Type C chopper.

This type of chopper works in first and second quadrant.

The Type C chopper is combination of both Type A and Type B chopper.

Type D Chopper

The load voltage is either positive or negative but load current always positive in the Type D chopper.

This chopper works in first as well as fourth quadrant.

<u>Type E Chopper</u>

The load voltage and load current are either positive or negative in the Type E chopper.

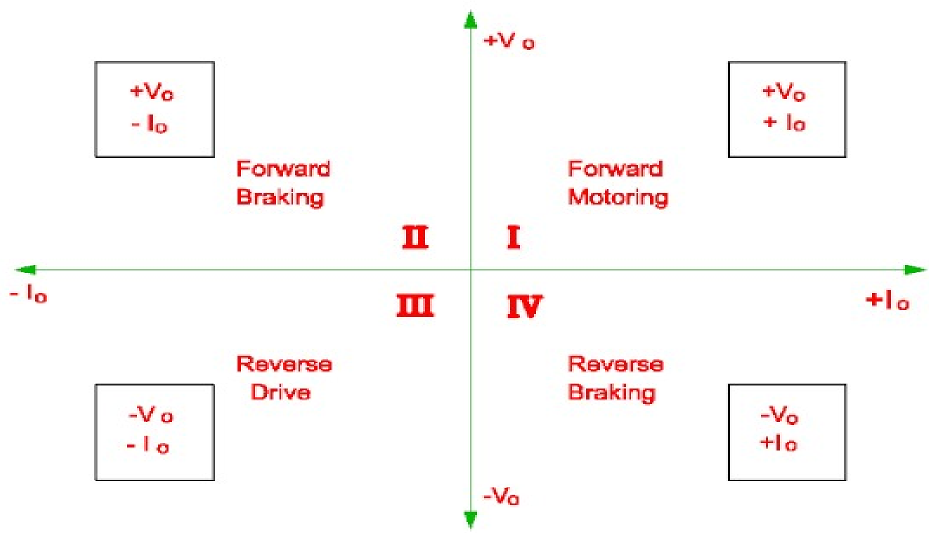
This type of chopper works in all quadrants. It is a combination of two Type C quadrants.

Forward Drive

If the chopper works in first quadrant, it is called as forward drive because the output voltage and current both are positive in this quadrant.

Reverse Drive

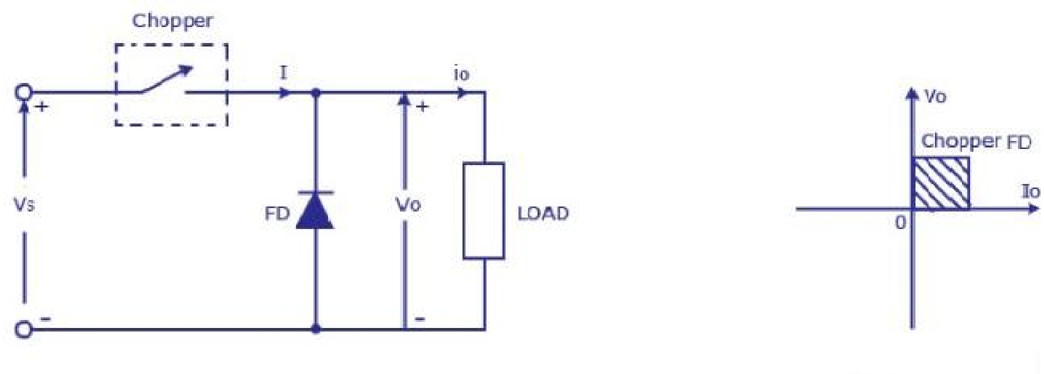
If the chopper works in third quadrant, it is called as reverse drive because the output voltage and current both are negative in this quadrant.
The DC motor works as DC generator in the second and fourth quadrant because the power flows from load to supply side in this quadrant.



CHOPPER CONFIGURATION

TYPE A, B, C, D AND E CHOPPER

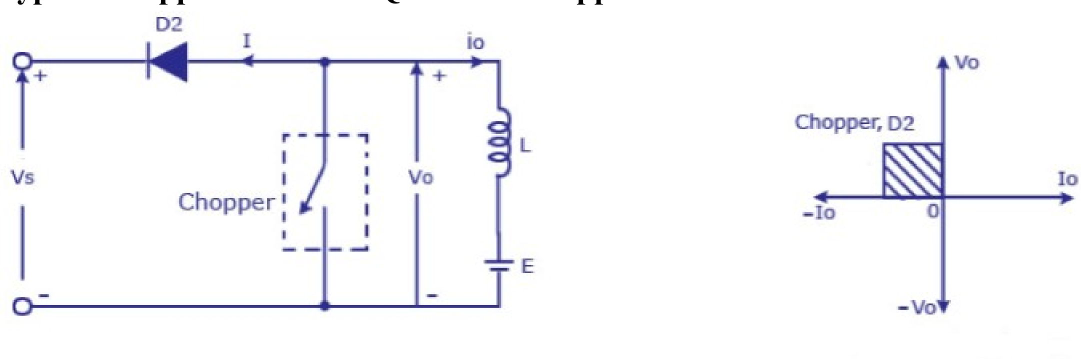
Type A Chopper or First–Quadrant Chopper
This type of chopper is shown in the figure. It is known as first-quadrant chopper or type A chopper. When the chopper is on, v0 = VS as a result and the current flows in the direction of the load. But when the chopper is off v0 is zero but I0 continues to flow in the same direction through the freewheeling diode FD, thus average value of voltage and current say V0 and I0 will be always positive as shown in the graph.



Chopper First Quadrant

In type A chopper the power flow will be always from source to the load. As the average voltage V_0 is less than the dc input voltage V_s

Type B Chopper or Second-Quadrant Chopper

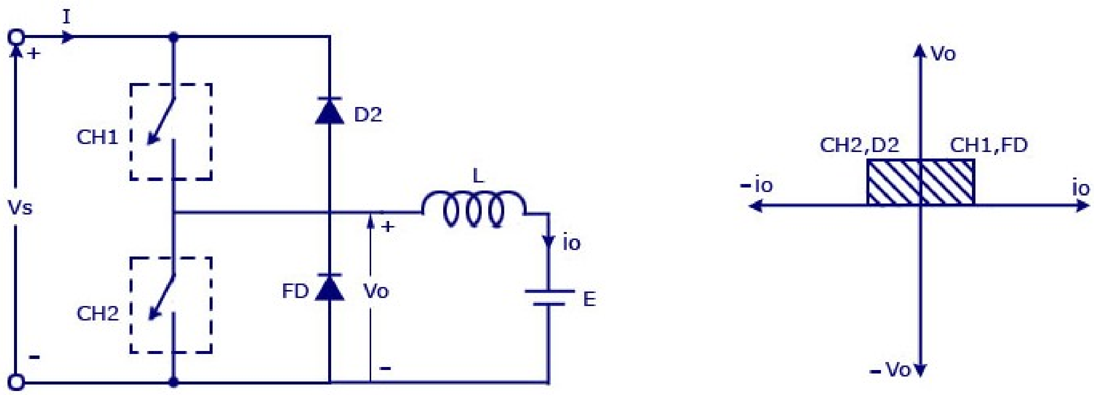


Chopper Second Quadrant

In type B or second quadrant chopper the load must always contain a dc source E. When the chopper is on, v_0 is zero but the load voltage E drives the current through the inductor L and the chopper, L stores the energy during the time T_{on} of the chopper. When the chopper is off, v_0 =(E+ L . *di/dt)* will be more than the source voltage Vs. Because of this the diode D2 will be forward biased and begins conducting and hence the power starts flowing to the source. No matter the chopper is on or off the current I_0 will be flowing out of the load and is treated negative. Since V_O is positive and the current I_0 is negative, the direction of power flow will be from load to source. The load voltage V_0 = (E+L *.di/dt)* will be more than the voltage V_s so the type B chopper is also known as a step up chopper .

Type -C chopper or Two-quadrant type-A Chopper

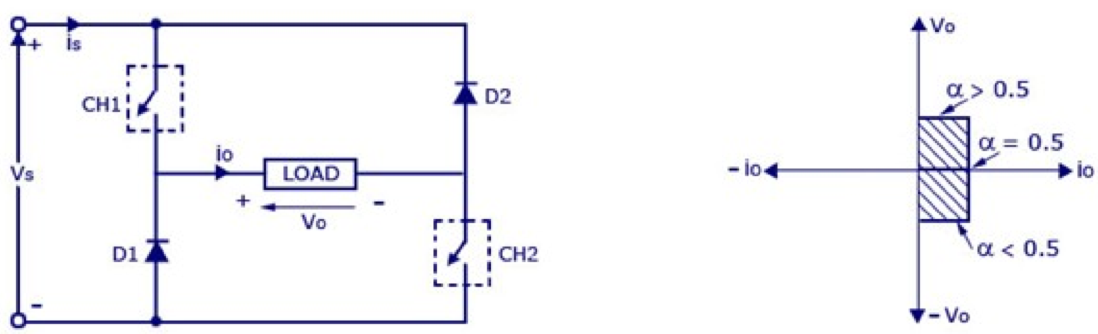
Type C chopper is obtained by connecting type –A and type –B choppers in parallel. We will always get a positive output voltage V_0 as the freewheeling diode FD is present across the load. When the chopper is on the freewheeling diode starts conducting and the output voltage v_0 will be equal to V_s . The direction of the load current i_0 will be reversed. The current i_0 will be flowing towards the source and it will be positive regardless the chopper is on or the FD conducts. The load current will be negative if the chopper is or the diode D2 conducts. We can say the chopper and FD operate together as type-A chopper in first quadrant. In the second quadrant, the chopper and D2 will operate together as type –B chopper.



Chopper Two Quadrant

The average voltage will be always positive but the average load current might be positive or negative. The power flow may be life the first quadrant operation ie from source to load or from load to source like the second quadrant operation. The two choppers should not be turned on simultaneously as the combined action my cause a short circuit in supply lines. For regenerative braking and motoring these type of chopper configuration is used.

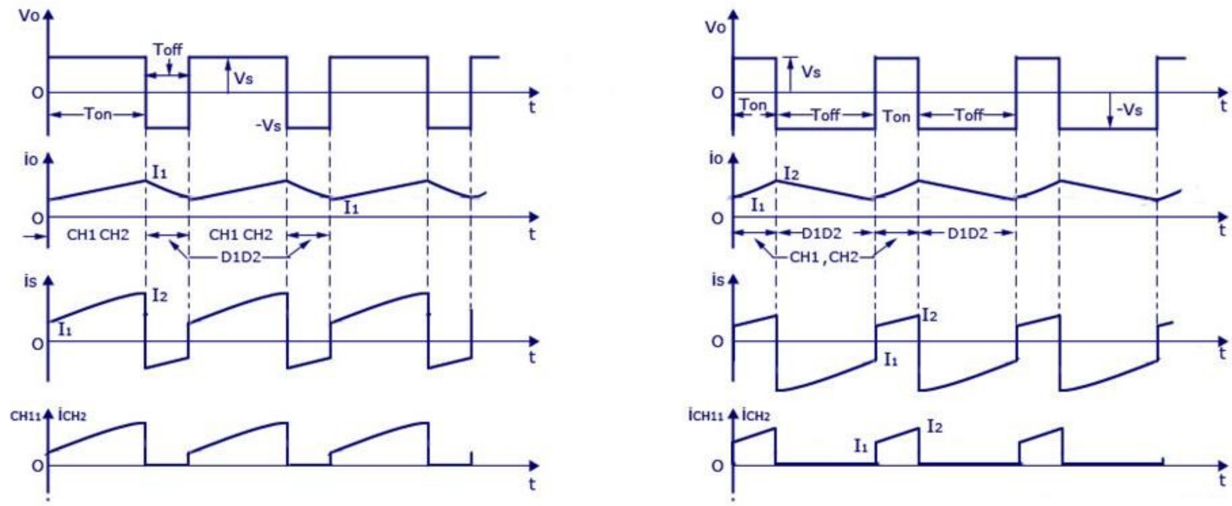
Type D Chopper or Two-Quadrant Type –B Chopper



Two Quadrant Type B-chopper or D-chopper Circuit

Two Quadrant Type B chopper or D Chopper Circuit

The circuit diagram of the type D chopper is shown in the above figure. When the two choppers are on the output voltage v_0 will be equal to V_s . When $v_0 = – V_s$ the two choppers will be off but both the diodes D1 and D2 will start conducting. V_0 the average output voltage will be positive when the choppers turn-on the time T_{on} will be more than the turn off time T_{off} it's shown in the wave form below. As the diodes and choppers conduct current only in one direction the direction of load current will be always positive.



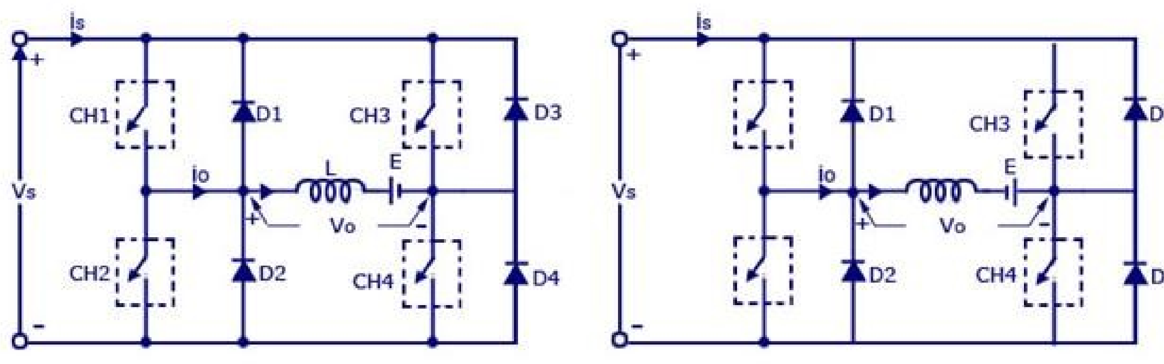
Positive First Quadrant Operation and Negative Fourth Quadrant Operation

The power flows from source to load as the average values of both v_0 and i_0 is positive. From the wave form it is seen that the average value of V_0 is positive thus the forth quadrant operation of type D chopper is obtained.

From the wave forms the Average value of output voltage is given by $V0= (V_s T_{on}-V_s T_{off})/ T = V_s.(T_{on}-T_{off})/T$

Type –E chopper or the Fourth-Quadrant Chopper

Type E or the fourth quadrant chopper consists of four semiconductor switches and four diodes arranged in antiparallel. The 4 choppers are numbered according to which quadrant they belong. Their operation will be in each quadrant and the corresponding chopper only be active in its quadrant.



E-type Chopper Circuit Diagram With Load emf E and E Reversed

- **First Quadrant**

During the first quadrant operation the chopper CH4 will be on. Chopper CH3 will be off and CH1 will be operated. AS the CH1 and CH4 is on the load voltage v_0 will be equal to the source voltage V_s and the load current i_0 will begin to flow. v_0 and i_0 will be positive as the first quadrant operation is taking place. As soon as the chopper CH1 is turned off, the positive current freewheels through CH4 and the diode D2. The type E chopper acts as a step- down chopper in the first quadrant.

- **Second Quadrant**

In this case the chopper CH2 will be operational and the other three are kept off. As CH2 is on negative current will starts flowing through the inductor L . CH2, E and D4. Energy is stored in the inductor L as the chopper CH2 is on. When CH2 is off the current will be fed back to the source through the diodes D1 and D4. Here (E+L.di/dt) will be more than the source voltage $V_{s.}$ In second quadrant the chopper will act as a step-up chopper as the power is fed back from load to source

- **Third Quadrant**

In third quadrant operation CH1 will be kept off , CH2 will be on and CH3 is operated. For this quadrant working the polarity of the load should be reversed. As the chopper CH3 is on, the load gets connected to the source V_s and v_0 and i_0 will be negative and the third quadrant operation will takes place. This chopper acts as a step-down chopper

- **Fourth Quadrant**

CH4 will be operated and CH1, CH2 and CH3 will be off. When the chopper CH4 is turned on positive current starts to flow through CH4, D2, E and the inductor L will store energy. As the CH4 is turned off the current is feedback to the source through the diodes D2 and D3, the operation will be in fourth quadrant as the load voltage is negative but the load current is positive. The chopper acts as a step up chopper as the power is fed back from load to source.

3. UNDERSTAND THE INVERTERS AND CYCLO-CONVERTERS

INVERTERS

A device that converts de power into ac power at desired output voltage and frequency is called an inverter. Some industrial applications of inverters are for adjustable-speed ac drives, induction heating, stand by air-craft power supplies, UPS (uninterruptible power supplies) for computers, hvdc transmission lines etc

CLASSIFY INVERTERS.

1. Inverters can be broadly classified into two types;

Voltage source inverters and current source inverters. A voltage-fed inverter (VFl), or voltage-source inverter (VSl), is one in which the dc source h small or negligible impedance. In other words, a voltage source inverter has stiff de voltage source at its input terminals.

A current-fed inverter (CFI) or current-source inverter (CSl) is fed with adjustable current from a dc source of high impedance, i.e. from a stiff dc current source. In: a CSI fed with stiff current source, output current waves are not affected by the load.

2. From the viewpoint of connections of semiconductor devices, inverters are classified as under:
   1. Bridge inverters 2. Series inverters 3. Parallel inverters
3. According to the Type of Load
   Single-phase Inverter
   Three-phase Inverter
4. According to the Number of Levels at the Output
   Regular Two-Level Inverter
   Multi-level Inverter

EXPLAIN THE WORKING OF SERIES INVERTER.

Series Inverter (Load Commutated Inverter or Self Commutated Inverter)

The commutating components L and C are connected in series with the load therefore this inverter is called as SERIES INVERTER.

The value of commutating components is selected such that the circuit becomes under damped.

The anode current itself becomes zero in this inverter resulting the SCR turns off automatically therefore this inverter is also called as SELF COMMUTATED OR LOAD COMMUTATED INVERTER.

Power Circuit Diagram

The power circuit diagram of the series inverter is shown in the figure A. The SCR T1 and SCR T2 are turned on at regular interval in order to achieve desirable output voltage and output frequency.

The SCR T2 is kept off at starting condition and polarity of voltage across capacitor is shown in the figure A.

The operation of the series inverter is explained as follows.

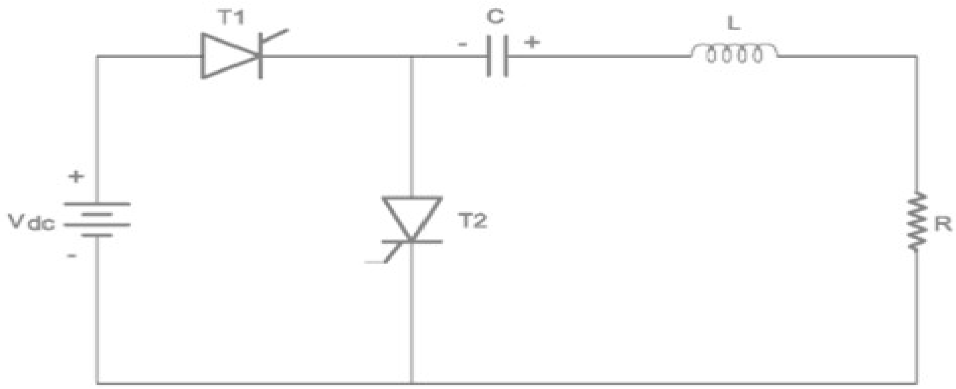


FIGURE A : BASIC SERIES INVERTER

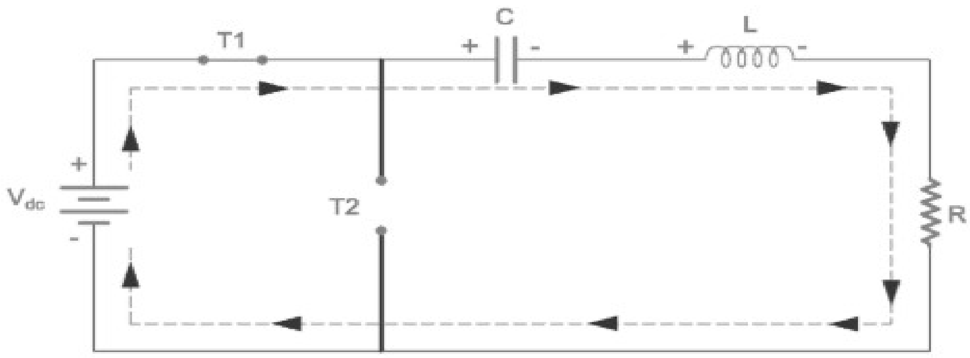


FIGURE B : EQUIVALENT CIRCUIT WHEN SCR T1 'ON'

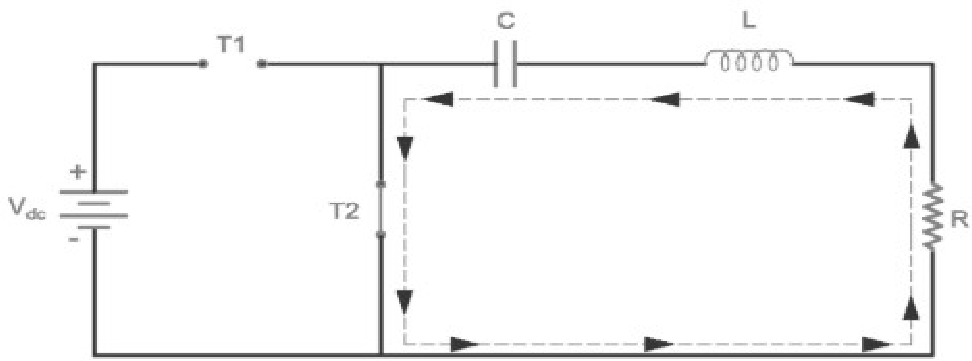


FIGURE C : EQUIVALENT CIRCUIT WHEN SCR T2 'ON'

Operation

Mode 1

The voltage V_{dc} directly applies to RLC series circuit as soon as the SCR T1 is turned on.

The polarity of capacitor charging is shown in the figure B.

The nature of the load current is alternating as there is under damped circuit of the commutating components.

The voltage across capacitor becomes + V_{dc} when the load current becomes maximum.

The voltage across capacitor becomes $+2V_{dc}$ when the load current becomes zero at point a (figure D).

The SCR T1 automatically turns off at point a because the load current becomes zero.

Mode 2

The load current becomes zero from point a to b as the SCR T1 turns off in this time period.
The SCR T1 and SCR T2 are turned off in this time duration and voltage across capacitor becomes equal to +2 V_{dc}.

Mode 3

The SCR T2 is turned on at point b due to it receives positive capacitor voltage.
The discharging of capacitor is done through SCR T2 and R – L circuit as shown in the figure C.
The load current becomes zero after it becomes maximum in the negative direction.
The capacitor discharges from +2 V_{dc} to – V_{dc} during this time and SCR T2 turns off automatically at point C due to load current becomes zero.
The SCR T2 turns off during point C to D and SCR T1 again turns on. This way cycle repeat after it complete one turns.
The positive AC output voltage half cycle generates due to DC voltage source whereas negative half cycle generates due to capacitor.
There is always some time delay kept between one SCR turned on time and other SCR turned on time.
The DC output gets short circuited due to continuous conduction of both SCRs if there is no time delay between SCRs.
The time duration ab and cd must be greater than the SCR specific turn off time and it is called as dead zone.

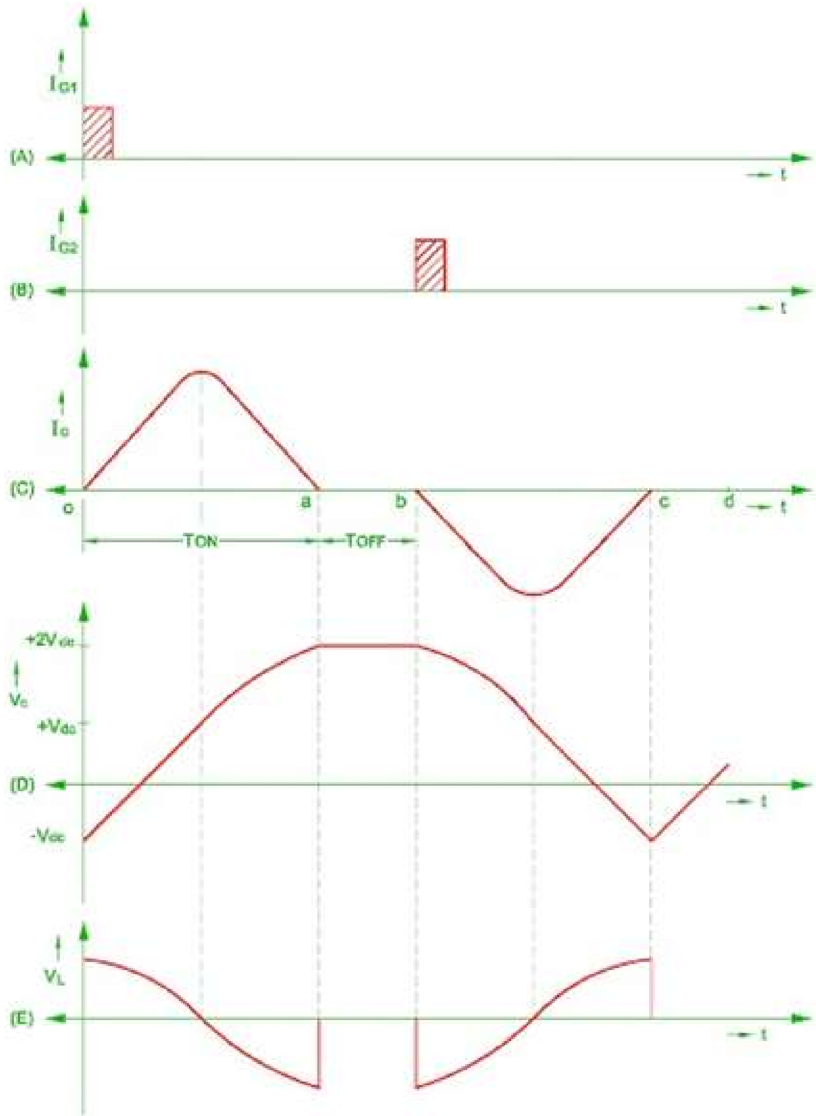


FIGURE D : VOLTAGE AND CURRENT WAVEFORMS OF SERIES INVERTER
(A) Gate pulse for SCR T1 (B) Gate pulse for SCR T2 (C) output current,
(D) Capacitor voltage (E) Load voltage

Limitations of Series Inverter

The limitation of series inverter is as given below.

The load current flows only during positive half cycle from supply source.

The DC supply source gets short circuited if SCR T1 and SCR T2 simultaneously turned on.

The rating of commutating components should high because the load current flows through it.

The load voltage waveform gets distorted if the dead zone time or SCR turns on time high.

The maximum output frequency of the inverter should be less than the ringing frequency.

The DC supply source is short circuited if the output frequency of the inverter is higher than the ringing frequency.

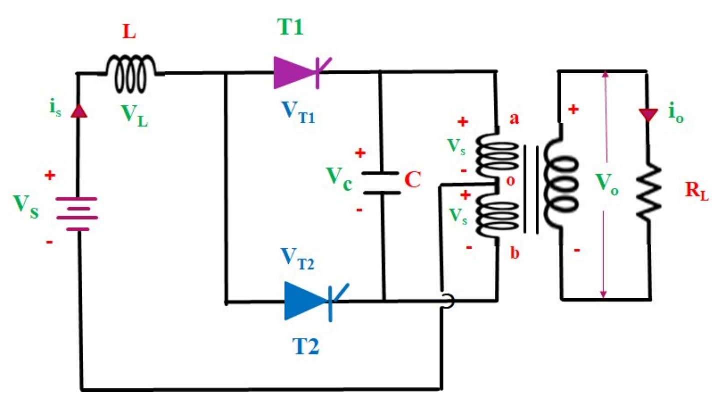
The maximum current during each high cycle and its time duration depends upon parameter of load this will result in poor regulation of the inverter output.

Applications

- This type of inverter generates sinusoidal waveform whose output frequency is in the range of 200 Hz to 100 kHz therefore it is applicable for
- Induction heating
- Sonar transmitter
- Fluorescent lighting and
- Ultrasonic generator

EXPLAIN THE WORKING OF PARALLEL INVERTER

Parallel inverter circuit consist of two thyristor T1 and T2, a transformer, inductor L and a commutating component C. Capacitor (C) is connected in parallel with the load via transformer therefore it is called a parallel inverter. And inductor (L) is connected in series with supply to make the source current constant. Here we also use a center -tapped transformer. Centre tapping is done in the primary winding of transformer so, primary winding is divided into two equal halves ao and ob.

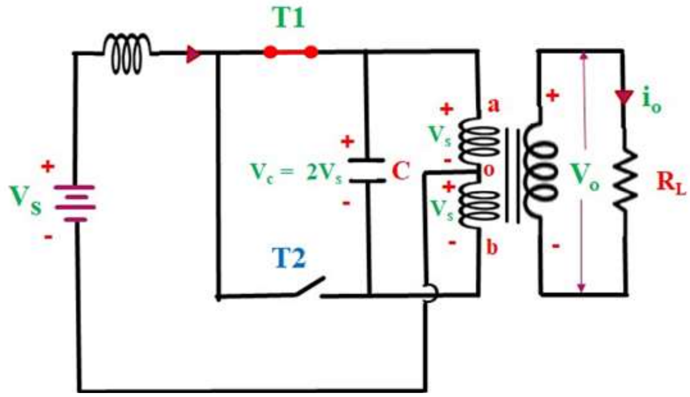


Circuit Diagram of Parallel Inverter

Operation of Parallel Inverter:

The operation is divided into four modes:

Mode I (0< t < t1): In this mode we give firing pulse to thyristor T1 and T1 get turned on and T2 is turned off. Current flow from Supply Vs …. T1…. ao (upper half of primary winding) …. back to Vs. As a result, Vs voltage is induced across upper as well as lower half of the primary winding of transformer. And Vs voltage is induced in secondary winding.

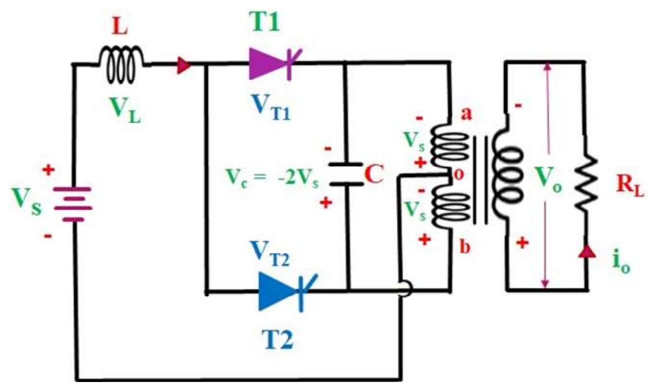


Mode I – Operation where T1 ON and T2 OFF

So, output voltage across load is Vs.

So, the total voltage across primary winding is 2Vs. Here capacitor is connected in parallel with primary winding therefore capacitor charge with 2Vs voltage with upper plate is positive and lower plate is negative.

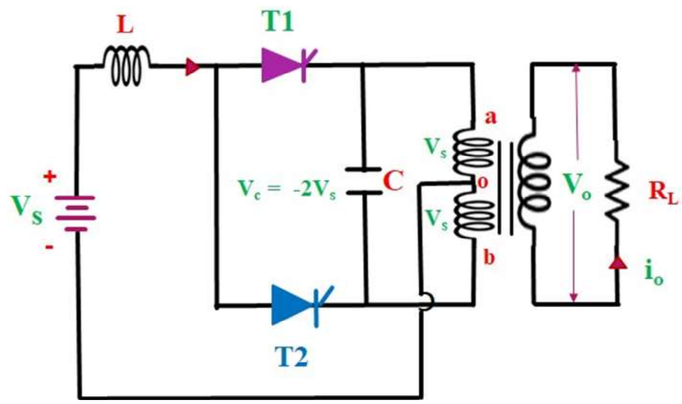
Mode II (t1< t < t3): In this duration we give firing pulse to thyristor T2 and T2 get turned on. At this time capacitor start discharging through T1 therefore T1 turned OFF. This time current flow from supply Vs …. T2…. bo (lower half of primary winding) …. back to Vs.



Mode II – Operation

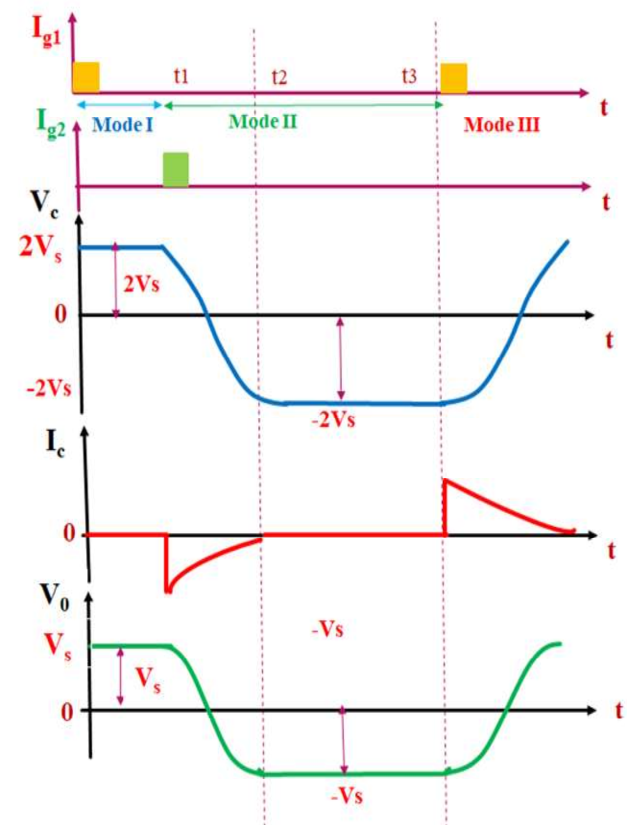
Now this time capacitor charged with upper plate is negative, from +2Vs at t=t1 to -2Vs at t=t2. Load voltage also changes from Vs at t=t1 to -Vs at t=t2. After t=t2 voltage across capacitor is maintain constant -2Vs between t= t2 to t3.So, load voltage is also constant -Vs.

Mode III (t3< t < t4): In this mode again, we give firing pulse to thyristor T1 and T1 get turned on. At this time capacitor start discharging through T2 therefore T2 turned OFF. This time current flow from supply Vs …. T1…. ao (upper half of primary winding) …. back to Vs. So, the total voltage across primary winding is 2Vs.



Mode III – Operation

Now this time capacitor charged with upper plate is positive, from -2Vs at t=t3 to +2Vs at t=t4. Load voltage also changes from Vs at t=t3 to -Vs at t=t4.So, output voltage across load is Vs.



Waveform of parallel Inverter 1) Ig1 is the gate current given to T1 2) Ig2 is the gate current given to T2. 3) Vc capacitor voltage 4) Ic current across capacitor 5) Vo output voltage waveform

EXPLAIN THE WORKING OF SINGLE-PHASE BRIDGE INVERTER.

Single phase half bridge inverter

Circuit Description:-

- Two Thyristor S1 and S2 are used along with two feedback diode D1 and D2 respectively.
- Resistive load is connect between point A and B, as shown in fig:-
- Supply voltage is divided into 2 parts, here two DC voltage source are used V/2 and V/2.
- Fig of the single phase half bridge inverter is given below:

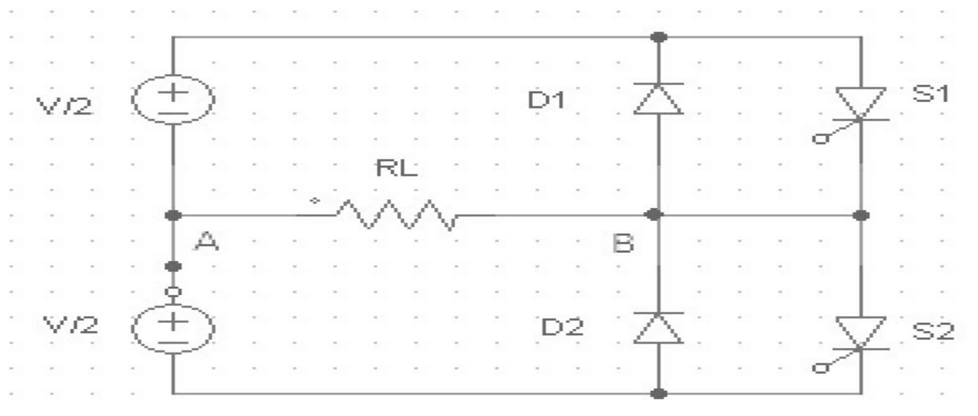


Fig : Single phase half bridge inverter

Working:

- Mode 1 (0 to T/2):-

- During this mode switch S1 is ON and switch S2 is OFF From period 0 to T/2.
- Current flowing path during this mode is V/2-S1-B-R(Load resistor)-A-V/2.
- Hence the voltage across the load is positive V/2

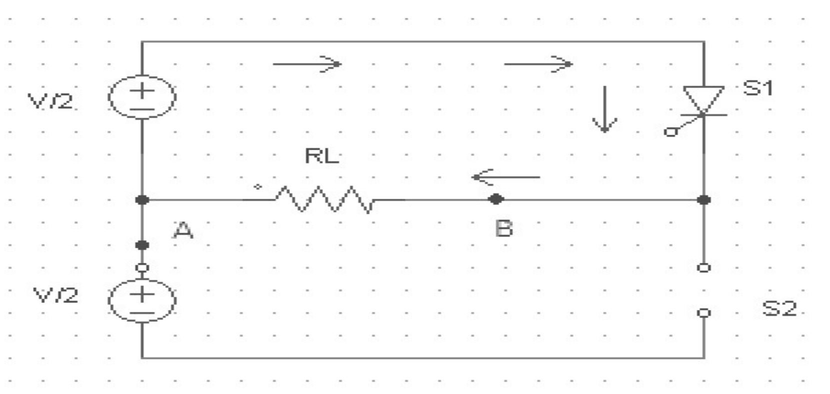


Fig: conducting mode 1

Mode 2 (T/2 to T):-

- During this mode switch S1 is OFF and switch S2 is ON From period T/2 to T.
- Current flowing path during this mode is V/2-A-R(Load resistor)-B-S2-V/2.
- Hence the voltage across the load is negative V/2.

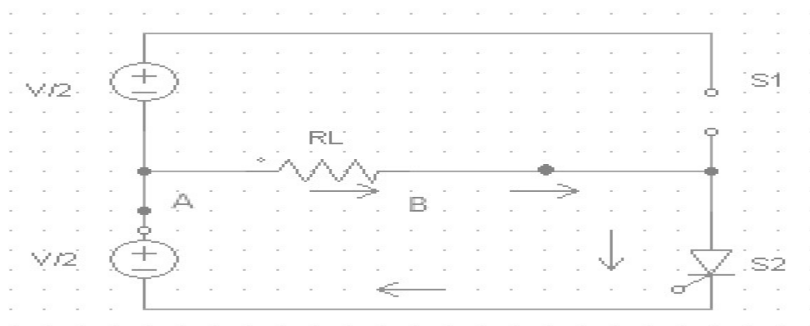


Fig: conducting mode 2

1. Load is resistive hence it does not store any charge therefore feedback diode D1 and D2 are not effective here.
2. The main drawback of half bridge inverter is that two DC voltage source are require. By using full bridge inverter we can overcome that drawback.

Waveform of output voltage thyristor current with resistive load are shown in fig

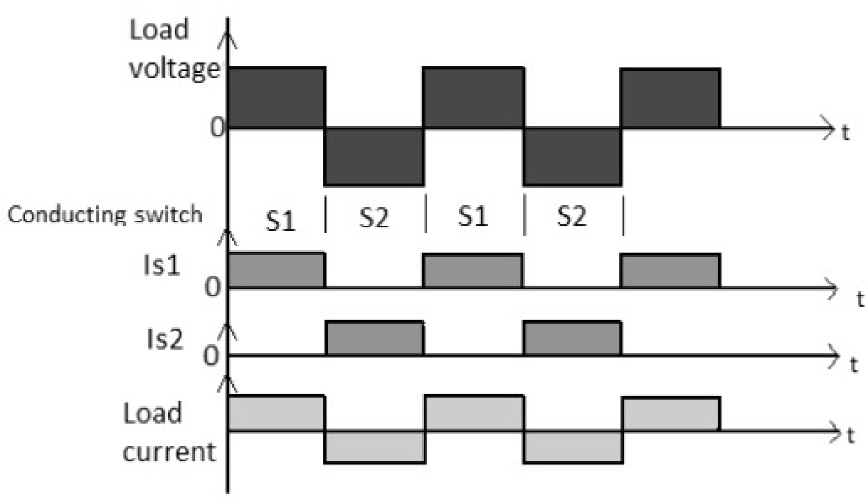


fig: waveform single phase half bridge inverter

Single phase Full bridge inverter

Circuit Description:-

- Four thyristor are used in full bridge inverter. Thyristor S1 and S2 are used along with two feedback diode D1 and D2 and thyristor S3 and S4 are used along with another two feedback diode D3 and D4 respectively.
- Resistive load is connect between point A and B,as shown in fig:-
- DC voltage source is applied to circuit.
- Fig of the single phase full bridge inverter is given below:

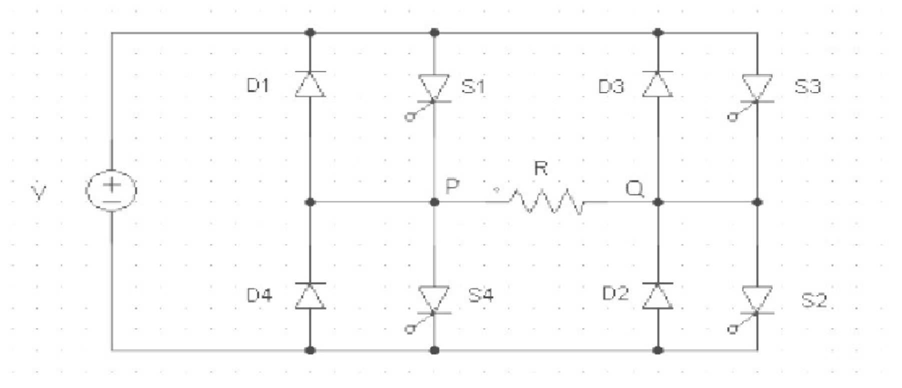
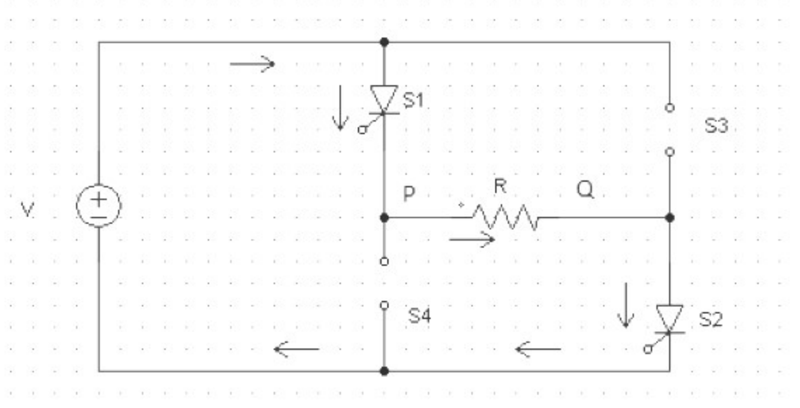


Fig: single phase full bridge inverter

Mode 1 (0 to T/2):-

- During this mode switch S1 and switch S2 are ON and switch S3 and switch S4 are OFF From period 0 to T/2.
- Current flowing path during this mode is Vdc – S1- P -R(load reistor) – Q – S2 – Vdc.
- Voltage across the load resistor is positive Vdc



fig; conducting mode 1

Mode 2 (T/2 to T):-

- During this mode switch S3 and switch S4 are ON and switch S1 and switch S2 are OFF From period T/2 to T.
- Current flowing path during this mode is Vdc – S3 – Q – R(load reistor) – P – S4 – Vdc.
- Voltage across the load resistor is negative Vdc.

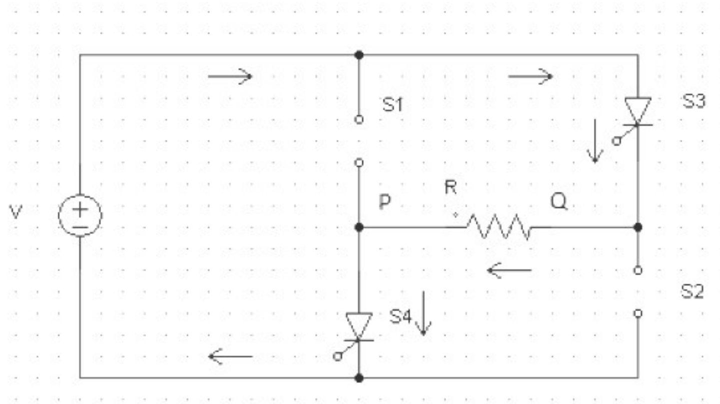


fig : conducting mode 2

1. Load is resistive hence it does not store any charge. therefore, feedback diode D1, D2, D3 and D4 are not effective here.

Waveform of output voltage thyristor current with resistive load are shown in fig:

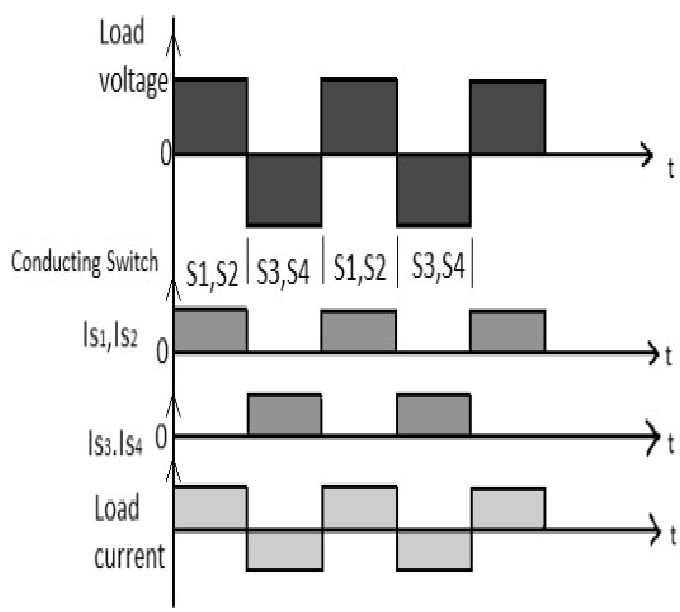


Fig: waveform single phase full bridge inverter

EXPLAIN THE BASIC PRINCIPLE OF CYCLO-CONVERTER

Cyclo-converters are AC to AC converters and are used to vary the frequency of a supply to a desired load frequency. A cycloconverter achieves this through synthesizing the output waveform from segments of the AC supply (without an intermediate DC link). These are naturally commutated, direct frequency converters that use naturally commutated thyristors. These are mainly used in high power applications up to tens of megawatts for frequency reduction.

Some of the applications of Cyclo-converter include high power AC drives, propulsion systems, high frequency induction heating, synchronous motors in sea and undersea vehicles, electromagnetic launchers, etc.

EXPLAIN THE WORKING OF SINGLE-PHASE STEP UP & STEP DOWN CYCLO-CONVERTER.

STEP-DOWN CYCLOCONVERTER

Step-down cycloconverter is a device which steps down the fixed frequency power supply input into some lower frequency. It is a frequency changer. If f_s & f_o are the supply and output frequency, then $f_o < f_s$ for this cycloconverter.

The most important feature of step-down cycloconverter is that it does not require force commutation. Line or Natural Commutation is used which is provided by the input AC supply.

Circuit Diagram:

There are two circuit configurations of a step-down cycloconverter: Mid-point and Bridge type**.** This article focuses on the mid-point type. The operation for continuous and discontinuous type of RL load is explained for mid-point type cycloconverter.

Figure below shows the circuit diagram of mid-point type cycloconverter. The positive direction of voltage and current are marked in the diagram.

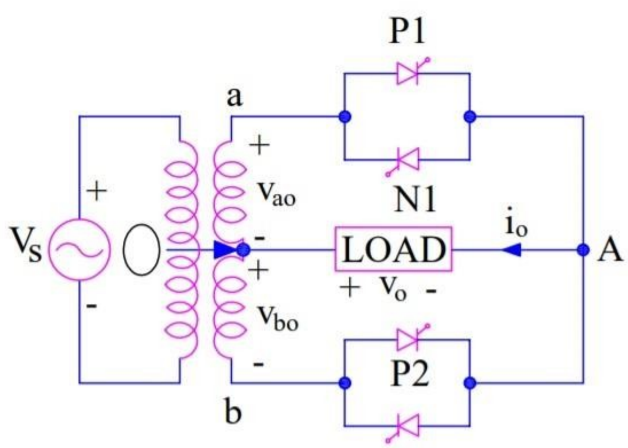


Fig. - 1

Working of Step-down Cycloconverter:

The working principle of step-down cycloconverter is explained for discontinuous and continuous load current. The load is assumed to be comprised of resistance (R) & inductance (L).

Discontinuous Load Current:

For positive cycle of input AC supply, the terminal A is positive with respect to point O. This makes SCRs P1 forward biased. The forward biased SCR P1 is triggered at ωt = 0. With this, load current i_o starts building up in the positive direction from A to O. Load current i_o becomes zero at ωt = β>π but less than (π+α). Refer figure-2. The thyristor P1 is thus, naturally commutated at ωt = β which is already reversed biased after π.

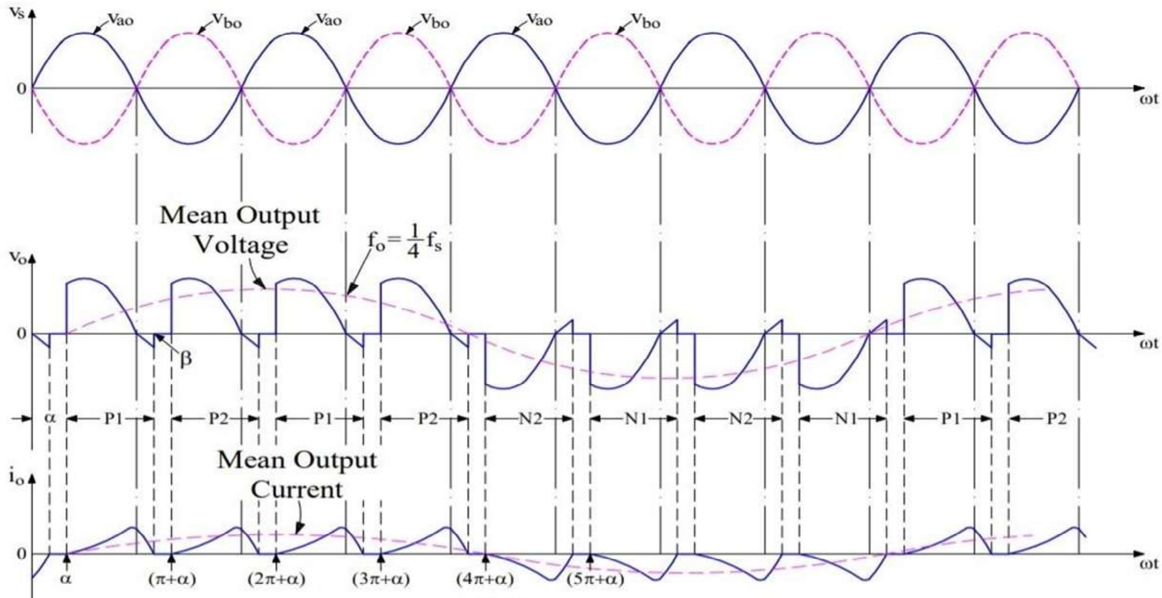


Fig.-2

After half a cycle, b is positive with respect to O. Now forward biased thyristor P2 is fired at ωt = (π+α). Load current is again positive from A to O and builds up from zero as shown in figure-2. At ωt = (π+ β), i_o decays to zero and P2 is naturally commutated. At ωt = (2π+α), P is again turned ON. Load current in figure-2 is seen to be discontinuous.

After four positive half cycles of load voltage and load current, thyristor N2 is gated at (4π+α) when O is positive with respect to b. As N2 is forward biased, it starts conducting but the direction of load current is reverse this time i.e. it flows from O to A. After N2 is triggered, O is positive with respect to "a" but before N1 is fired, i_o decays to zero and N2 is naturally commutated. Now when N1 is gated at (5π+α), i_o again builds up but it decays to zero before thyristor N2 in sequence is again gated.

In this manner, four negative half cycles of load voltage and load current, equal to number of positive half cycles of load voltage & current, are generated. Now P1 is again triggered to fabricate four positive half cycles of load voltage and so on. It may be noted that, natural commutation is achieved for discontinuous current load.

Form figure-2, the waveform of mean load voltage & current may be noted. It is clear that the output frequency of load voltage & current is (¼) times of input supply frequency.

STEP-UP CYCLOCONVERTER

Working Principle of Step-up Cycloconverter:

The working principle of a step-up cycloconverter is based on switching of thyristors in a proper sequence. The thyristor acts as a power switch. These switches are arranged is a specific pattern

so that the output power is available for both the positive and negative half of the input power supply. Forced commutation technique is used to turn OFF the conducting thyristor.
Two circuit configurations are possible for step-up cycloconverter: Mid-point Type and Bridge Type**.** In this article, we will consider mid-point type of circuit arrangement for better understanding of working principle.
Circuit Diagram:
Figure below shows the circuit diagram of Mid-point step-up cycloconverter:

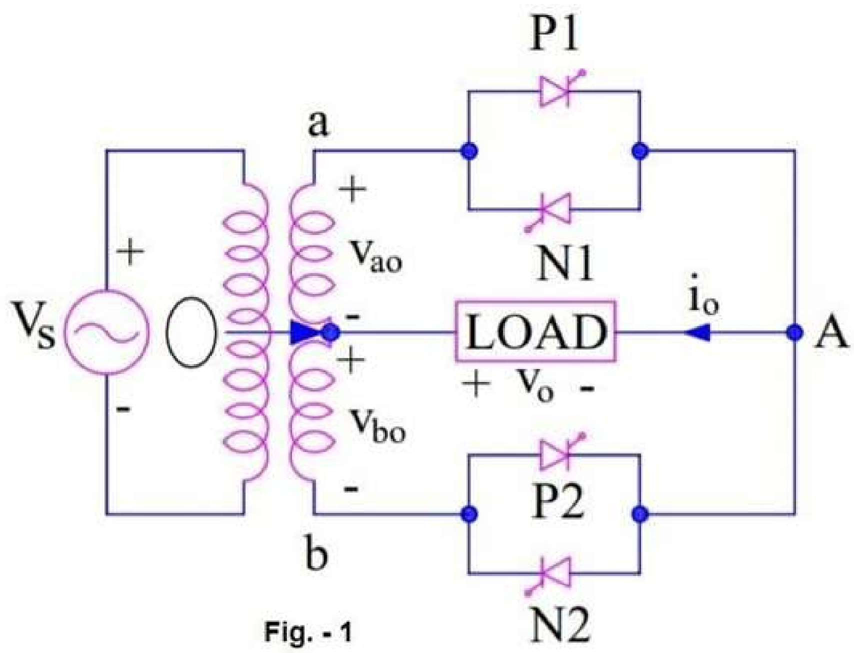


Fig. - 1

The circuit consists of a single phase transformer with mid tap on the secondary winding and four thyristors. Two of these thyristors P1 & P2 are for positive group. Here positive group means when either P1 or P2 conducts, the load voltage is positive. Other two thyristors N1 & N2 are for negative group. Load is connected between secondary winding mid-point O and terminal A. The load is assumed resistive for simplicity. Assumed positive direction for voltage and current are marked in the circuit diagram.

Operation of Step-up Cycloconverter:
During the positive half cycle of input supply voltage, positive group thyristors P1 & N2 are forward biased for $\omega t = 0$ to $\omega t = \pi$. As such SCR P1 is fired to turn it ON at $\omega t = 0$ such that load voltage is positive with terminal A positive and O negative. The load voltage, thus, follows the positive envelop of the input supply voltage. At some time instant $\omega t = \omega t_1$, the conducting thyristor P1 is force commutated and the forward biased thyristor N2 is fired to turn it ON. During the period N2 conducts, the load voltage is negative because O is positive & A is negative this time. The load or output voltage traces the negative envelop of the supply voltage. This is shown in figure below.

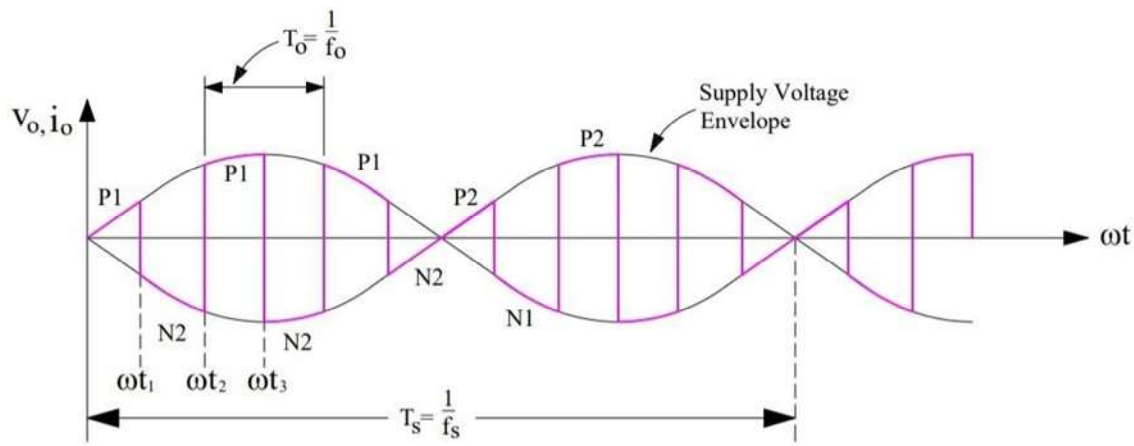


Fig. - 2

At $\omega t = \omega t_2$, N2 is force commutated and P1 is turned ON. The load voltage is now positive and follows the positive envelop of the supply voltage. At $\omega t = \pi$, terminal "b" is positive with respect to terminal "a"; both SCRs P2 & N1 are therefore forward biased from $\omega t = \pi$ to $\omega t = 2\pi$. AT $\omega t = \pi$, N2 is force commutated and forward biased SCR P2 is turned ON. The load voltage is positive and follows the positive envelop of supply voltage.
If the supply frequency is f_s and output frequency is f_o, P2 will be force commutated at $\omega t = (1/2f_s) + (1/2f_o)$. Carefully note this from the waveform shown in the figure-2.
When P2 is force commutated, forward biased SCR N1 is turned ON. This time, the load voltage is negative and follows the negative envelop of the supply input.
In this manner, SCRs P1, N2 for the first half cycle; P2, N1 in the second half cycle and so on are switched alternately between positive and negative envelops at a high frequency. This results in output frequency f_o more than the input supply frequency f_s. In our example of figure-2, note that there is a total of 6 cycles of output in one cycle of input supply. This means that frequency of output voltage is 6 times of input frequency i.e. $f_o = 6f_s$.

APPLICATIONS OF CYCLO-CONVERTER.

Application of Cycloconverter:
The general use of cycloconverter is to provide either a variable frequency power from a fixed input frequency power as in AC motor speed control or a fixed frequency power from a variable frequency power as in aircraft or wind generators.
Some of the major applications of cycloconverter are as follows:

- Speed control of high power AC drives
- Induction heating
- Static VAR compensation
- It is used for converting variable speed alternator voltage to constant frequency output voltage for use as power supply in aircrafts or shipboards.

4. UNDERSTAND APPLICATIONS OF POWER ELECTRONIC CIRCUITS

LIST APPLICATIONS OF POWER ELECTRONIC CIRCUITS.

- Our Daily Life: If we look around ourselves, we can find a whole lot of power electronics applications such as a fan regulator, light dimmer, air-conditioning, induction cooking, emergency lights, personal computers, vacuum cleaners, UPS (uninterrupted power system), battery charges, etc.
- Automotives and Traction: Subways, hybrid electric vehicles, trolley, fork-lifts, and many more. A modern car itself has so many components where power electronic is used such as ignition switch, windshield wiper control, adaptive front lighting, interior lighting, electric power steering and so on. Besides power electronics are extensively used in modern traction systems and ships.
- Industries: Almost all the motors employed in the industries are controlled by power electronic drives, for eg. Rolling mills, textile mills, cement mills, compressors, pumps, fans, blowers, elevators, rotary kilns etc. Other applications include welding, arc furnace, cranes, heating applications, emergency power systems, construction machinery, excavators etc.
- Defense and Aerospace: Power supplies in aircraft, satellites, space shuttles, advance control in missiles, unmanned vehicles and other defense equipments.
- Renewable Energy: Generation systems such as solar, wind etc. needs power conditioning systems, storage systems and conversion systems in order to become usable. For example solar cells generate DC power and for general application we need AC power and hence power electronic converter is used.
- Utility System: HVDC transmission, VAR compensation (SVC), static circuit breakers, generator excitation systems, FACTS, smart grids, etc.

LIST THE FACTORS AFFECTING THE SPEED OF DC MOTORS.

According to the speed equation of a d.c. motor we can write,

$$N \propto \frac{E_b}{\phi}$$

or $$N = K\frac{(V - I_a R)}{\phi} \text{ r.p.m.} \qquad \text{(i)}$$

where $R = R_a$ for shunt motor

$= R_a + R_{se}$. for series motor

The factors affecting DC control are therefore:

- The applied voltage
- The flux

- The voltage across an armature

Considering these factors, speed control can then be achieved through the following techniques:

- Flux control method: This is done by varying the current via the field winding, thus altering the flux.
- Rheostat control: changing the armature route resistance which also changes the applied voltage across the armature.
- Voltage method: changing the applied voltage

SPEED CONTROL FOR DC SHUNT MOTOR USING CONVERTER.

Speed Control of DC Shunt Motor

The power circuit diagram for speed control of the DC Shunt Motor is shown in the figure E.

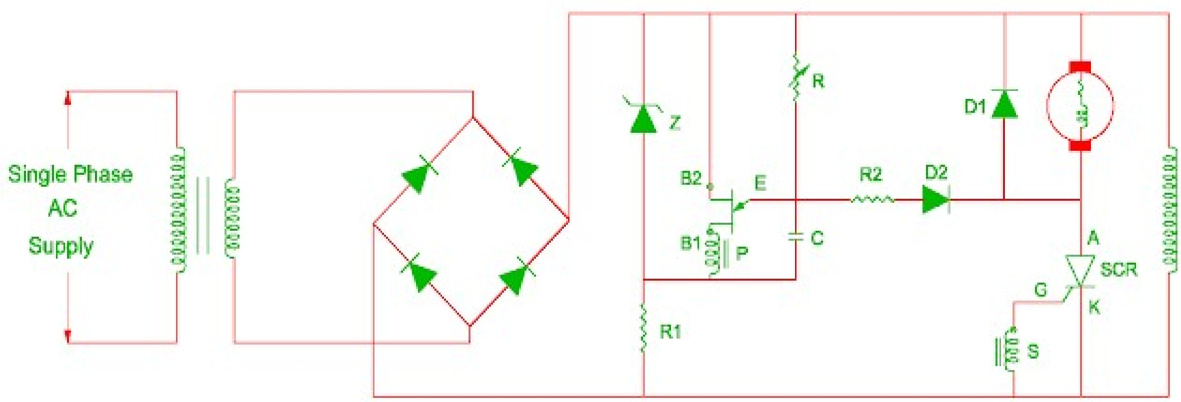


FIG E : Speed control of DC Shunt Motor

1. The function of the bridge rectifier is to converter alternating voltage in to direct voltage.
2. The zener diode clips the voltage and provides constant voltage.
3. The field winding of the DC shunt motor is connected across supply voltage.
4. The SCR is connected in series with the armature winding of the DC shunt motor.
5. The charging of capacitor is done through variable resistor R.
6. When the voltage across capacitor becomes equal to peak point voltage, the UJT turns on.
7. The discharging of capacitor is done through path C – EB1 – Primary of pulse transformer – C.
8. As soon as the pulse transformer primary energies, the SCR gets pulse through pulse transformer secondary.
9. Now the current passes through the armature winding of the DC motor.
10. The charging rate of the capacitor depends upon variable resistor R.
11. If the value of variable R is set minimum, the charging of capacitor done faster resulting UJT turns on in short time.

12. This will resulting the firing angle of the SCR decreases and DC motor speed increases.
13. If the value of resistor R set at maximum, the firing angle of SCR increases and DC motor speed decreases.
14. As the field winding gets constant voltage, the motor speed is directly proportional to back emf.
15. If the armature winding drop is neglected, the speed of the DC motor is directly proportional to armature voltage.
16. The speed of the DC motor is adjusted by the firing angle of the SCR.
17. When the UJT turns on, the diode D2 forward biases and diode D1 reverse biased therefore the charging of capacitor is done only through variable resistor R.
18. The diode D1 reverse biases when current passes through the armature winding.
19. As soon as the current passes through the armature winding becomes zero, the stored energy of armature winding dissipates through diode D1.

Speed Regulation

1. As the motor speed increases, the back emf also increases and diode D2 becomes forward biased in this condition.
2. As the charging path of capacitor and resistor R2 becomes parallel, the charging current of capacitor decreases and firing angle of SCR increases.
3. This will result the speed of motor decreases. If the motor speed decreases by any chance, the back emf decreases.
4. This will result in small current passes through shunt resistor R2 and firing angle of SCR decreases due to charging rate of capacitor increases.
5. The DC shunt motor speed increases due to decrease of firing angle.

SPEED CONTROL FOR DC SHUNT MOTOR USING CHOPPER.

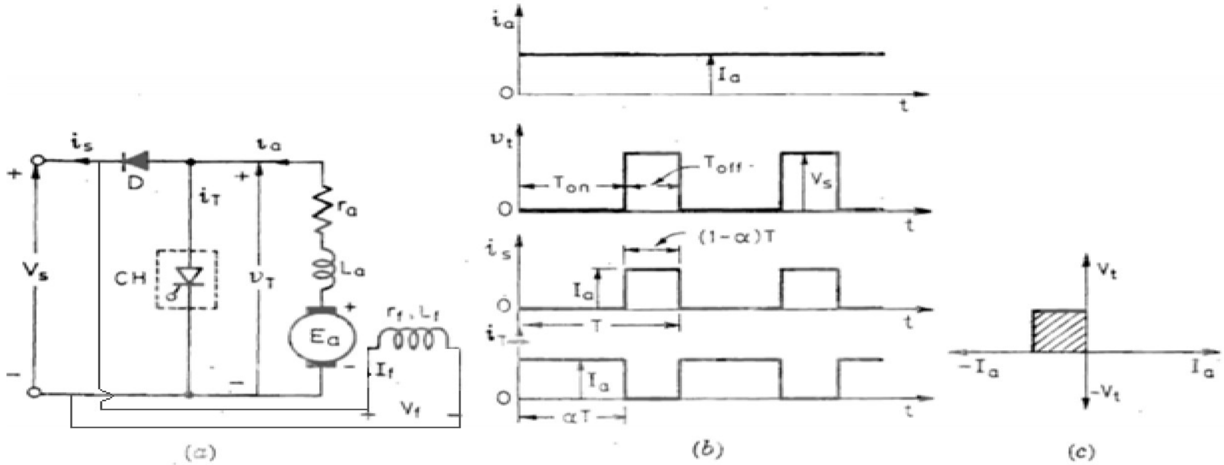
Chopper drive used for single quadrant regenerative braking control

In regenerative-braking control, the motor acts as a generator and the kinetic energy of the motor and connected load is returned to the supply.

During motoring mode, armature current la $=\frac{vt-Ea}{r_a}$, i.e. armature current is positive and the motor consumes power. In case load drives the motor at a speed such that average value of motor counter emf Ea ($=k_m.\omega_m$) exceeds Vt, Ia is reversed and power is delivered to the dc bus. The motor is then working as a generator in the regenerative braking mode.

The principle of regenerative braking mode is explained with the help of Fig.(a), where a separately-excited dc motor and a chopper are shown. For active loads, such as a train going down the hill or a descending hoist, let it be assumed that motor counter emf Ea is more than the source voltage Vs When chopper CH is on, current through armature inductance La rises as the armature terminals get short 'circuited through CH. Also, v_t =0 during Ton. When chopper is turned off, Ea being more than source voltage Vs diode D conducts and the energy stored in armature inductance is transferred to the source. During T_{off} $v_t = V_s$ On the assumption of continuous and ripple free armature current, the relevant voltage and current waveforms are shown in Fig. (b). Regenerative braking control offers second quadrant operation as armature terminal voltage has the same polarity but the direction of armature current is reversed, Figs. (a) and (c). From the waveforms of Fig.(b), the following relations can be derived: The average voltage across chopper (or armature terminals) is

$$\mathbf{V_t = \frac{T_o}{T} \cdot V_s = (1\text{-}\alpha)\ Vs}$$



Regenerative braking of separately excited dc motor (a) circuit diagram (b)waveform (c)quadrant operation

Power generated by the motor

$$= V_t \cdot I_a = (1 - \alpha)\, V_s \cdot I_a$$

Motor emf generated, $\quad E_a = K_m\, \omega_m = V_t + I_a\, r_a$

$$= (1 - \alpha)\, V_s + I_a\, r_a$$

Motor speed during regenerative braking,

$$\omega_m = \frac{(1 - \alpha)\, V_s + I_a r_a}{K_m}$$

Motor speed during regenerative braking,

$$\omega_m = \frac{(1 - \alpha)\, V_s + I_a r_a}{K_m}$$

When chopper is on, $E_a - I_a r_a - L_a \dfrac{di_a}{dt} = 0$

or $\quad (E_a - I_a r_a) = L_a \cdot \dfrac{di_a}{dt}$

With chopper on, L_a must store energy and current must rise, *i.e.* $\frac{di_a}{dt}$ must be positive or

$$(E_a - I_a r_a) \geq 0$$

When chopper is off, $E_a - I_a r_a - L_a \cdot \frac{di_a}{dt} = V_s$

or
$$V_s - (E_a - I_a r_a) = -L_a \cdot \frac{di_a}{dt}$$

With chopper off, $(E_a - I_a r_a)$ must be more than V_s for regeneration purposes and therefore $[V_s - (E_a - I_a r_a)]$ must be negative. This is possible only if current decreases during off period, *i.e.* $\frac{di_a}{dt}$ in the above expression must be negative.

$$\therefore \quad [V_s - (E_a - I_a r_a)] \leq 0$$
$$-(E_a - I_a r_a) \leq (-V_s)$$

or
$$(E_a - I_a r_a) \geq V_s$$

Eqs. can be combined to give the conditions for controlling the power during regenerative braking as

$$0 \leq (E_a - I_a r_a) \geq V_s$$

Eq. gives the conditions for the two voltages and their polarity for the regenerative braking control of dc separately-excited motor.

Minimum braking speed is obtained when $E_a - I_a r_a = 0$

or
$$K_m \, \omega_{mn} = I_a r_a$$

$\therefore$ Minimum braking speed $\qquad \omega_{mn} = \frac{I_a r_a}{K_m}$

Maximum possible braking speed is obtained when

$$E_a - I_a r_a = V_s$$

$\therefore$ Maximum braking speed, $\qquad \omega_{mx} = \frac{V_s + I_a r_a}{K_m}$

Thus regenerative braking control is effective only when motor speed is less than ω_{mx} and more than ω_{mn}. This can be expressed as

$$\omega_{mn} < \omega_m < \omega_{mx}$$
$$\frac{I_a r_a}{K_m} < \omega_m < \frac{V_s + I_a r_a}{K_m}$$

Therefore, the speed range for regenerative braking is $\frac{V_s + I_a r_a}{K_m} : \frac{I_a r_a}{K_m}$ or $(V_s + I_a r_a) : I_a r_a$.

Regenerative braking of chopper-fed separately-excited or self excited dc shunt motor is more stable, it is therefore discussed here. DC series motors, however, offer unstable operating characteristics during regenerative braking. As such, regenerative braking of chopper-controlled series motors is difficult.

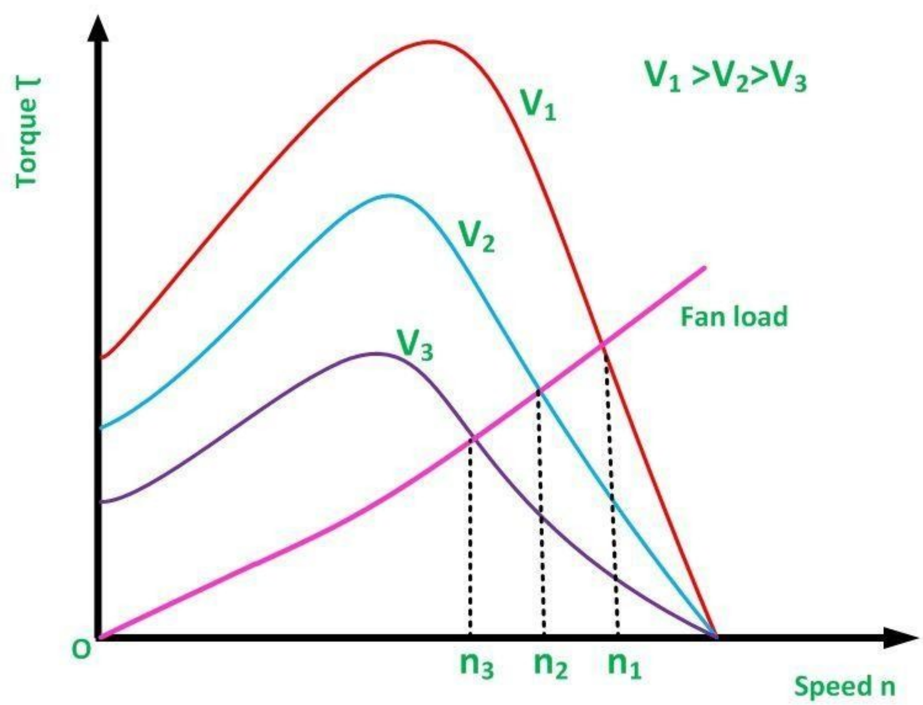
LIST THE FACTORS AFFECTING SPEED OF THE AC MOTORS

The factors that affect speed of the ac motors are

- Voltage
- Frequency
- Slip
- Current
- Rotor resistance

SPEED CONTROL OF INDUCTION MOTOR BY USING AC VOLTAGE REGULATOR.

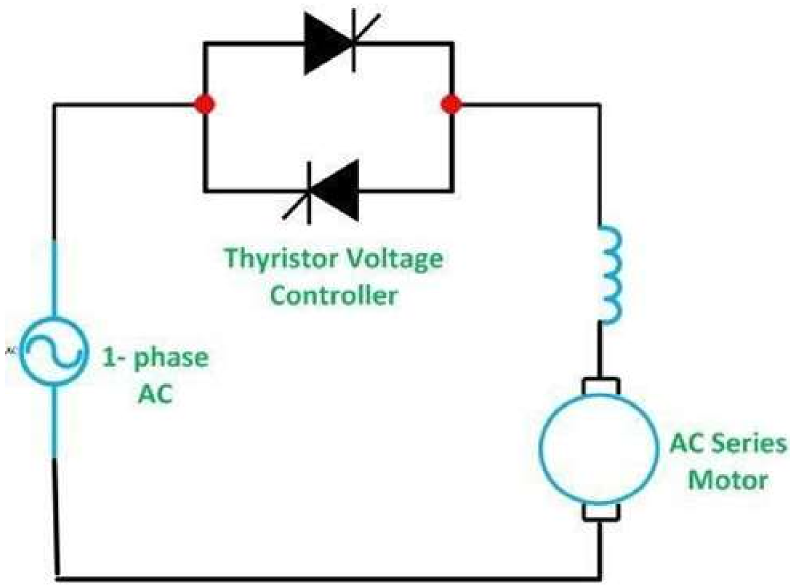
The speed of a three-phase induction motor can be varied by varying the supply voltage. As we already know that the torque developed is proportional to the square of the supply voltage and the slip at the maximum torque is independent of the supply voltage. The variation in the supply voltage does not alter the synchronous speed of the motor. The Torque-Speed Characteristics of the three-phase induction motors for varying supply voltage and also for fan load are shown below:



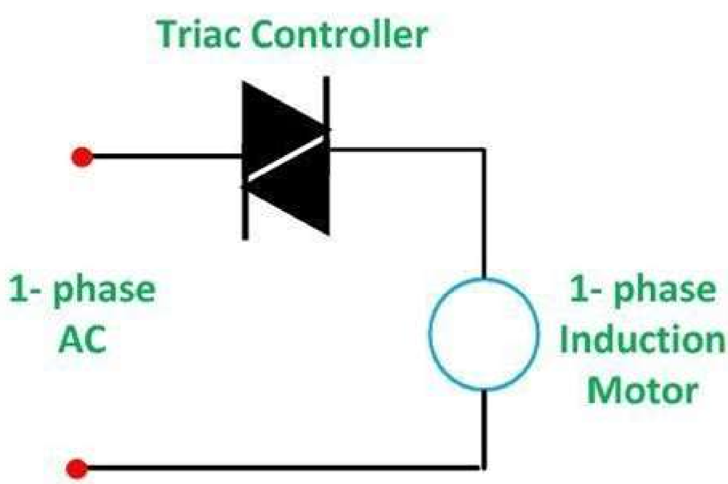
By varying the supplying voltage, the speed can be controlled. The voltage is varied until the torque required by the load is developed, at the desired speed. The torque developed is proportional to the square of the supply voltage and the current is proportional to the voltage.

Hence, to reduce the speed for the same value of the same current, the value of the voltage is reduced and as a result, the torque developed by the motor is reduced. This stator voltage control method is suitable for applications where the load torque decreases with the speed.

Thyristor voltage controller method is preferred for varying the voltage. For a single phase supply, two thyristors are connected back to back as shown in the figure below:

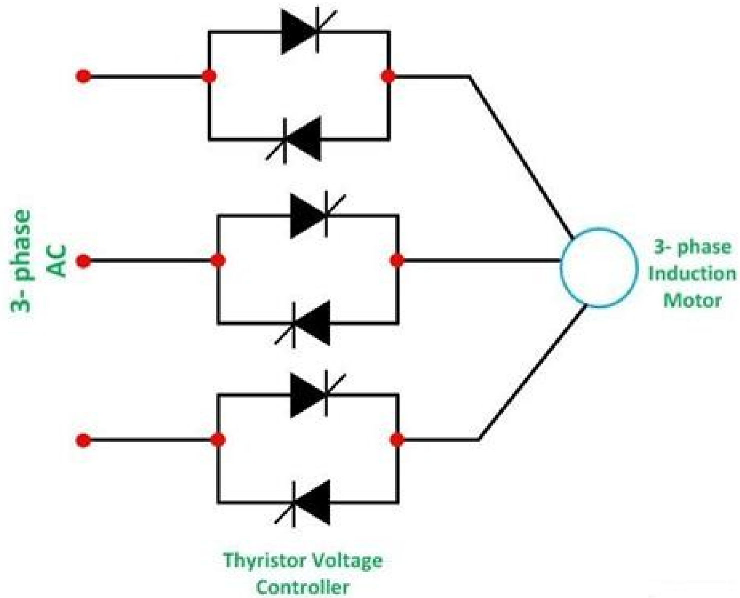


The domestic fan motors, which are single-phase, are controlled by a single-phase Triac Voltage Controller as shown in the figure below:



Speed control is obtained by varying the firing angle of the Triac. These controllers are known as Solid State fan regulators. As the solid-state regulators are more compact and efficient as compared to the conventional variable regulator. Thus, they are preferred over the normal regulator.

In the case of a three-phase induction motor, three pairs of thyristors are required which are connected back to back. Each pair consists of two thyristors. The diagram below shows the Stator Voltage Control of the three-phase induction motors by Thyristor Voltage Controller.



Each pair of the thyristor controls the voltage of the phase to which it is connected. Speed control is obtained by varying the conduction period of the Thyristor. For lower power ratings, the back-to-back thyristor pairs connected in each phase are replaced by the Traic.

Speed control of induction motor by using converters and inverters (V/F control).

For a 3-ph IM stator voltage per phase is given by

$v_i = \sqrt{2}\pi f_1 . N_{ph1} . \varphi . k_{w1}$

It is seen from above equation that if the ratio of supply voltage V_1 to supply frequency f_1 is kept constant, the air-gap flux remains constant. From Fig., the starting torque is given by

$$T_{e.st} = \frac{3}{\omega_s} \cdot \frac{V_1^2}{(r_1 + r_2)^2 + (x_1 + x_2)^2} \cdot r_2$$

As $(r_1 + r_2) << (x_1 + x_2)$ and $\omega_s = \frac{2\omega_1}{P}$, we get

$$T_{e \cdot st} = \frac{3P}{2\omega_1} \cdot \frac{V_1^2 \cdot r_2}{\omega_1^2 (l_1 + l_2)^2}$$

$$= \frac{3P}{2\omega_1} \cdot \left(\frac{V_1}{\omega_1}\right)^2 \cdot \frac{r_2}{(l_1 + l_2)^2}$$

maximum torque is given by

$$T_{e.m} = \frac{3}{\omega_s} \cdot \frac{V_1^2}{2(x_1 + x_2)}$$

$$= \frac{3P}{2\omega_1} \cdot \frac{V_1^2}{2 \cdot \omega_1 (l_1 + l_2)}$$

$$= \frac{3P}{4} \cdot \left(\frac{V_1}{\omega_1}\right)^2 \frac{1}{l_1 + l_2}$$

The Eq. shows that if $\frac{v1}{\omega 1}$, or air-gap flux φ, is kept constant, the maximum torque remains unaltered. Eq. indicates that starting torque inversely proportional to supply frequency $\omega_1$even if air-gap flux is kept constant. At low values of frequencies, the effect of resistances cannot be neglected as compared to the reactances. This has the effect of reducing the magnitude of maximum torque at lower frequencies as shown in Fig. In practice, at low frequencies, the supply voltage is increased to maintain the level of maximum torque. This method of speed control is also called volts / hertz control.

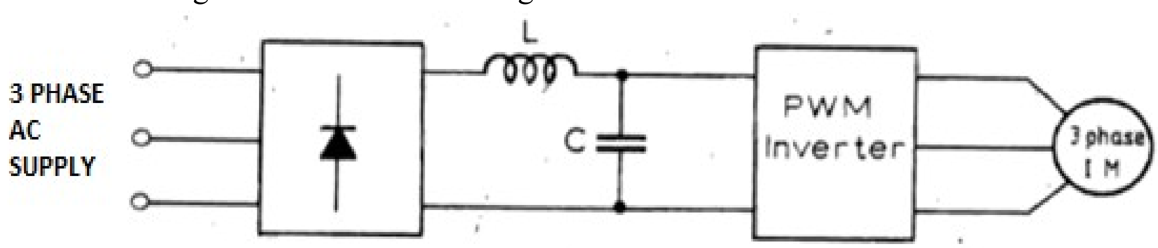
If stator reactance is neglected the slip at which maximum torque occurs is given by

$$s_m = \frac{r_2}{x_1 + x_2}$$

$$= \frac{r_2}{\omega_1 (l_1 + l_2)}$$

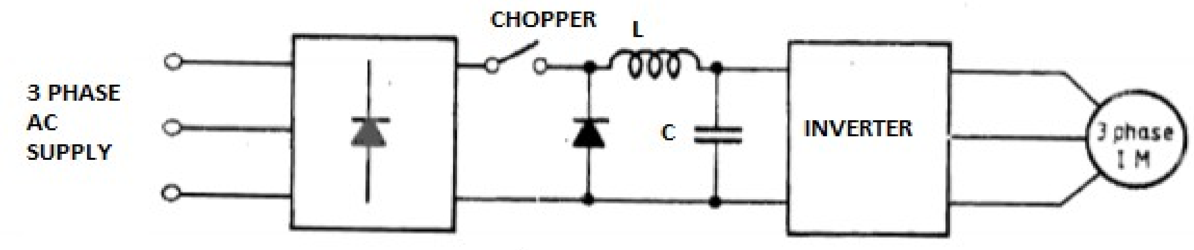
As the supply frequency (ω_1 is reduced, the slip at maximum torque increases. In Fig.load torque T_L for a certain load is also shown. It is seen from this figure that as both voltage and frequency are varied (usually below their rated values), speed of the drive can be controlled. The control of both voltage and frequency can be carried out (so as to keep $\frac{V}{f}$ constant through the use of three-phase inverters or cycloconverters. Inverters are used in low and medium power drives whereas cycloconverters are suitable for high-power drives like cement mills, locomotives etc.

Variable voltage and variable -frequency can be obtained from voltage-source inverters. Four such circuit configurations are shown in Fig.



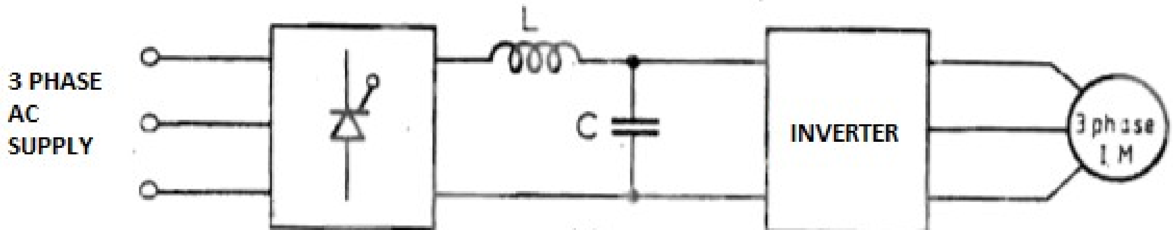
(a)

In Fig. Three-phase ac is converted to constant dc by diode rectifier. Voltage and frequency are both varied by PWM inverter. The circuitry between the rectifier and the inverter consists of an inductor L and capacitor C, called filter circuit. The function of filter circuit is to smooth dc input voltage to the inverter. This Circuitry in between rectifier and inverter is called dc link. In Fig. -regeneration is not possible because of diode rectifier. Also, inverter would inject harmonics into the 3-phase ac supply.



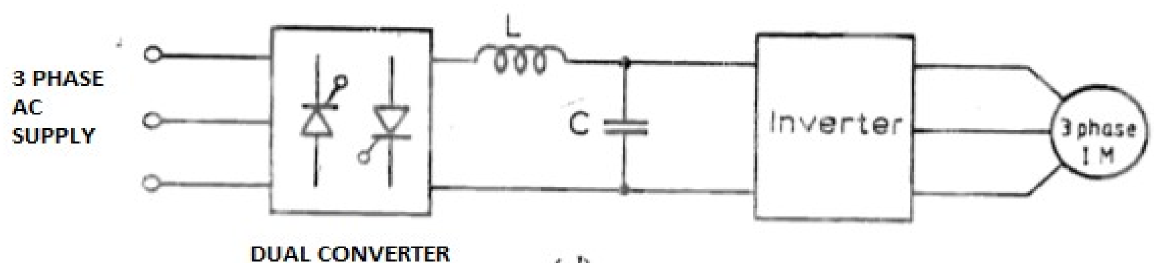
(b)

In Fig. three-phase ac is converted to dc by diode rectifier. Chopper varies the dc input voltage to the inverter and frequency is controlled by the inverter. Use of chopper reduces the harmonic injection into the ac supply. Regeneration is not feasible in the scheme of Fig.



(c)

Fig. uses a 3-phase controlled rectifier, dc link consisting of Land C and a force-commutated VSI. Voltage is regulated by controlled rectifier and frequency is varied within the inverter. Here regeneration is possible if three-phase full converter is used. Regeneration is also feasible in the scheme shown in Fig. below It uses a 3-phase dual -- converter, L-C filter and inverter. Level of dc input voltage to the inverter is regulated in dual converter whereas frequency is varied within the VSI inverter.



(d)

It may be observed from above that volt/hertz control offers speed control from standstill up to rated speed of IM. This method is similar to the armature-voltage control method used for the speed control of a dc motor.

WORKING OF UPS WITH BLOCK DIAGRAM.

Uninterruptible Power Supply | UPS

In a UPS, the energy is generally stored in flywheels, batteries, or super capacitors. When compared to other immediate power supply system, UPS have the advantage of immediate protection against the input power interruptions. It has very short on-battery run time; however this time is enough to safely shut down the connected apparatus (computers, telecommunication equipment etc) or to switch on a standby power source.

UPS can be used as a protective device for some hardware which can cause serious damage or loss with a sudden power disruption. Uninterruptible power source, Battery backup and Flywheel back up are the other names often used for UPS. The available size of UPS units ranges from 200 VA which is used for a solo computer to several large units up to 46 MVA.

Major Roles of UPS

When there is any failure in main power source, the UPS will supply the power for a short time. This is the prime role of UPS. In addition to that, it can also able to correct some general power problems related to utility services in varying degrees. The problems that can be corrected are voltage spike (sustained over voltage), Noise, Quick reduction in input voltage, Harmonic distortion and the instability of frequency in mains.

Types of UPS

Generally, the UPS system is categorized into On-line UPS, Off- line UPS and Line interactive UPS. Other designs include Standby on-line hybrid, Standby-Ferro, Delta conversion On-Line.

Off-line UPS

This UPS is also called as Standby UPS system which can give only the most basic features. Here, the primary source is the filtered AC mains (shown in solid path in figure 1). When the power breakage occurs, the transfer switch will select the backup source (shown in dashed path in figure 1). Thus we can clearly see that the stand by system will start working only when there is any failure in mains. In this system, the AC voltage is first rectified and stored in the storage battery connected to the rectifier.

When power breakage occurs, this DC voltage is converted to AC voltage by means of a power inverter, and is transferred to the load connected to it. This is the least expensive UPS system and it provides surge protection in addition to back up. The transfer time can be about 25 milliseconds which can be related to the time taken by the UPS system to detect the utility voltage that is lost.

The block diagram is shown below.

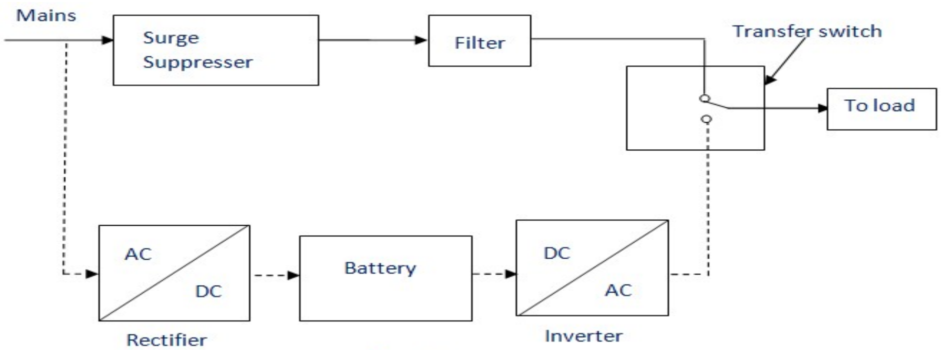


Figure 1

On-line UPS

In this **type of UPS**, double conversion method is used. Here, first the AC input is converted into DC by rectifying process for storing it in the rechargeable battery. This DC is converted into AC by the process of inversion and given to the load or equipment which it is connected (figure 2). This type of UPS is used where electrical isolation is mandatory. This system is a bit more costly due to the design of constantly running converters and cooling systems. Here, the rectifier which is powered with the normal AC current is directly driving the inverter. Hence it is also known as Double conversion UPS. The block diagram is shown below.

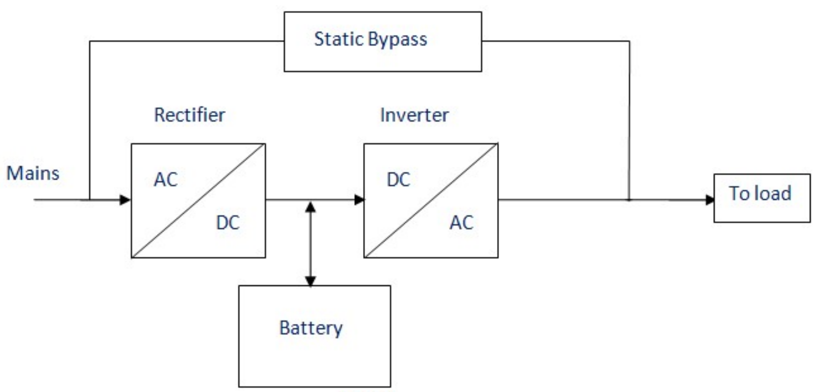


Figure 2

When there is any power failure, the rectifier have no role in the circuit and the steady power stored in the batteries which is connected to the inverter is given to the load by means of transfer switch. Once the power is restored, the rectifier begins to charge the batteries. To prevent the batteries from overheating due to the high power rectifier, the charging current is limited. During a main power breakdown, this UPS system operates with zero transfer time. The reason is that the backup source acts as a primary source and not the main AC input. But the presence of inrush current and large load step current can result in a transfer time of about 4-6 milliseconds in this system.

Applications of a UPS:

- Data Centers
- Industries
- Telecommunications

- Hospitals
- Banks and insurance
- Some special projects (events)

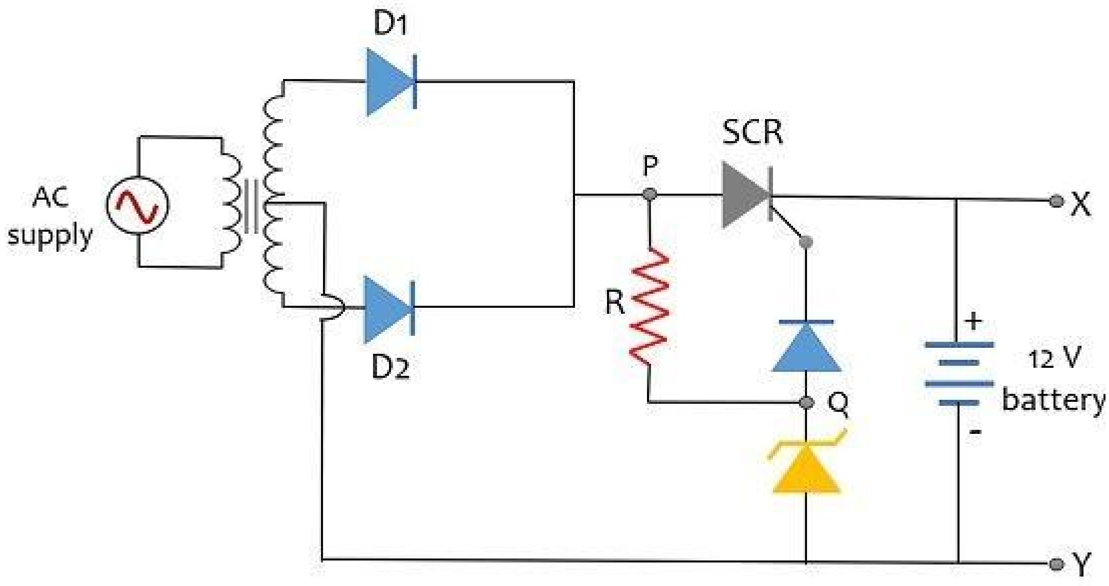
BATTERY CHARGER CIRCUIT USING SCR WITH THE HELP OF A DIAGRAM

An SCR-based battery charger makes use of the switching principle of the thyristor in order to get the specific output. The circuit includes a transformer, rectifier, and control circuit as its major elements.

As we have already discussed in the beginning that a small amount of ac or dc input voltage is needed for the purpose of charging the battery. So, the elements of the circuit help to provide the desired voltage to charge the battery.

Working of Battery Charger circuit using SCR

The figure below represents the circuit of a battery charger incorporating an SCR:



Battery Charger Circuit using SCR

Here, an ac voltage signal of value 230 V, 50 Hz is applied as input and the load is a 12 V battery that is required to be charged.

Following are the circuit elements:

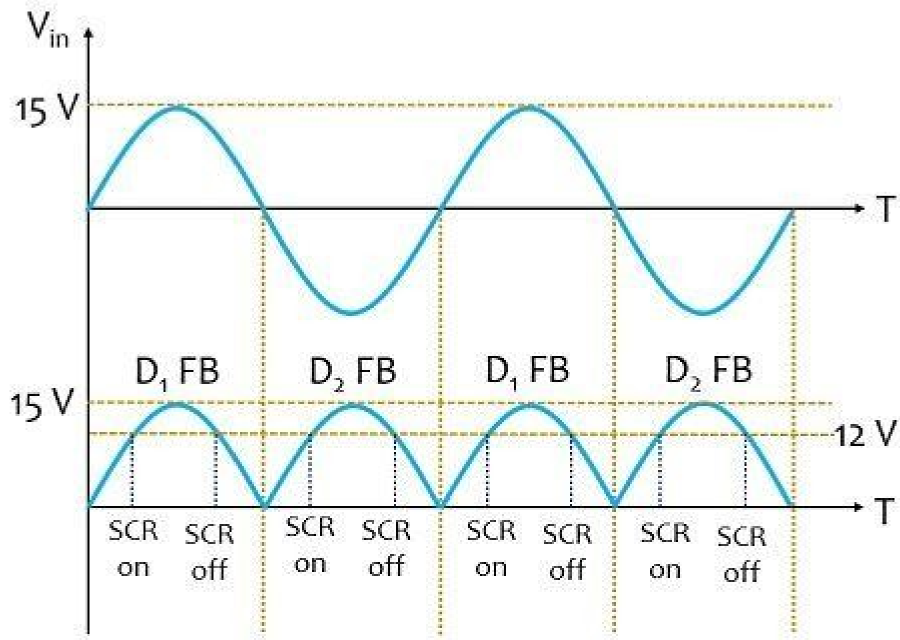
- AC supply
- Step down transformer
- Rectifier circuit
- SCR
- Zener diode as a voltage regulator
- Battery to be charged

Let us now understand how the above-given circuit operates.

So, initially, 230 V ac supply is provided to the step-down transformer which converts the high voltage given at the input of the primary winding into a low voltage which is obtained at the output of the secondary winding. So, here the voltage obtained at the other side of the transformer is 15V with respect to neutral.

From the circuit, it is clearly shown that the transformer forms connection with the rectifier circuit, hence the output of the transformer will be provided to the rectifier circuit. As we have an ac input signal, so let us understand how the rectifier circuit operates when the two halves of the ac signal are applied.

Initially, when the positive half of the ac input signal is applied then the diode D_1 in the above-given configuration will be forward biased and will conduct however, D_2 will be in reverse biased condition thus will not conduct. Conversely, when the negative half of ac input is applied then D_1 will not conduct but D_2 will be in conducting state, this is clearly shown in the waveform representation given below:



So, the rectifier circuit will provide rectified output i.e., the dc voltage at terminal P.

Here we have used a Zener diode with breakdown voltage of 10 V as a voltage regulator to regulate the voltage level of the circuit. Therefore, terminal Q will be at 10 V due to the presence of the Zener diode.

As the terminal voltage at P which is nothing but the rectified voltage is comparatively more than at terminal Q thus, this makes the SCR forward biased, allowing it to conduct and due to this current will start flowing through the load i.e., the **12 V battery**. And we have already discussed in the beginning that when current flows through the battery then the cells present within it stores the energy. In this way, the battery gets charged.

However, in case, the rectified voltage is less than that of terminal voltage at Q then automatically SCR will come in a reverse-biased state, getting it turned off and no flow of current through the battery will take place further.

Thus, we can say that here the SCR is acting as a switch that controls the voltage fed to the battery. Now, the question arises, once the battery gets fully charged how the circuit will operate**.** So, basically what happens within the circuit is as the rectified voltage here is 15 V, so once the battery gets fully charged (suppose it reaches 14.5 V) then the remaining value of the voltage at terminal P will be insufficient to cause further conduction through the SCR because now the rectified voltage will be less than the voltage at terminal Q. This will not allow current to reach the battery further and resultantly the charging circuit will get deactivated.

Basically, this comparison between rectified voltage and the charging potential is made using a comparator circuit. Once the charging potential falls below a certain value then the charging circuit will automatically get activated and again the charging of the battery will take place.

It is to be noted here that the value of the breakdown voltage of the Zener diode and the transformer in the circuit depends on the charging potential of the battery. Thus, the potential at which the battery will be charged will decide the value of these two circuit parameters.

Drawbacks of Battery charger circuit using SCR

This charging is a quite time taking process.

The rectifier circuit for ac to dc conversion, do not remove ac ripples as the filter circuit is absent here.

The process of charging and discharging is slow due to the presence of a half-wave rectifier.

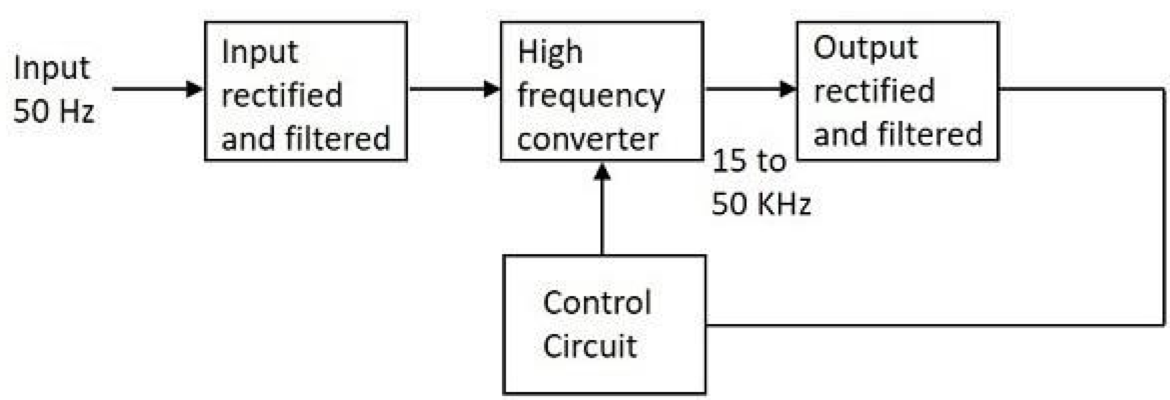
This is only suitable for charging the batteries with a low to medium ampere-hour rating.

BASIC SWITCHED MODE POWER SUPPLY (SMPS) - EXPLAIN ITS WORKING & APPLICATIONS

A switched-mode power supply (switched power supply,) is an electronic power supply that incorporates a switching regulator to convert electrical power efficiently.

Working

The working of SMPS can be understood by the following figure.



Let us try to understand what happens at each stage of SMPS circuit.

Input Stage

The AC input supply signal 50 Hz is given directly to the rectifier and filter circuit combination without using any transformer. This output will have many variations and the capacitance value of the capacitor should be higher to handle the input fluctuations. This unregulated dc is given to the central switching section of SMPS.

Switching Section

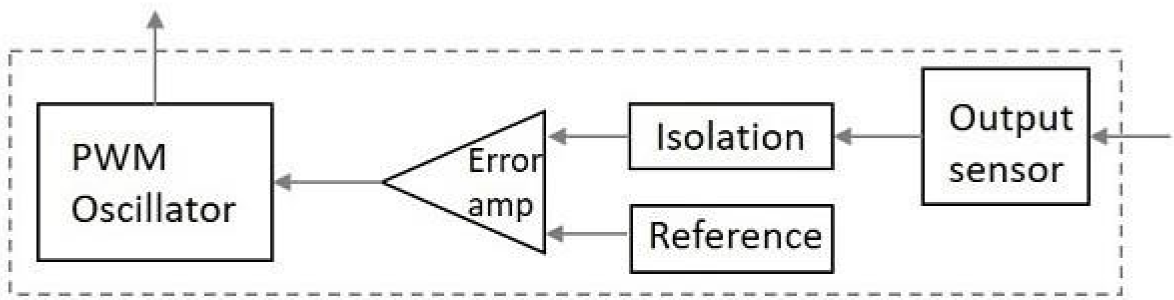
A fast switching device such as a Power transistor or a MOSFET is employed in this section, which switches ON and OFF according to the variations and this output is given to the primary of the transformer present in this section. The transformer used here are much smaller and lighter ones unlike the ones used for 60 Hz supply. These are much efficient and hence the power conversion ratio is higher.

Output Stage

The output signal from the switching section is again rectified and filtered, to get the required DC voltage. This is a regulated output voltage which is then given to the control circuit, which is a feedback circuit. The final output is obtained after considering the feedback signal.

Control Unit

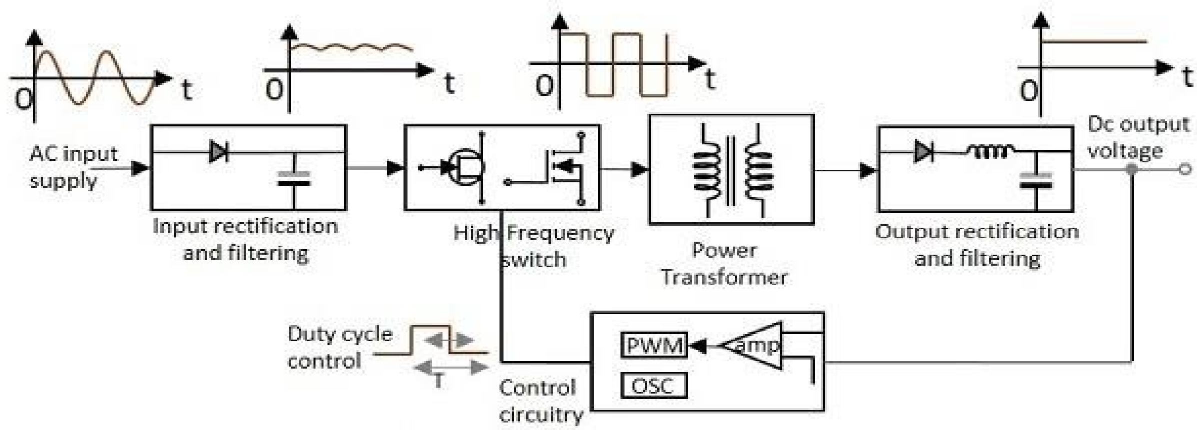
This unit is the feedback circuit which has many sections. Let us have a clear understanding about this from The following figure.



The above figure explains the inner parts of a control unit. The output sensor senses the signal and joins it to the control unit. The signal is isolated from the other section so that any sudden spikes should not affect the circuitry. A reference voltage is given as one input along with the signal to the error amplifier which is a comparator that compares the signal with the required signal level.

By controlling the chopping frequency the final voltage level is maintained. This is controlled by comparing the inputs given to the error amplifier, whose output helps to decide whether to increase or decrease the chopping frequency. The PWM oscillator produces a standard PWM wave fixed frequency.

We can get a better idea on the complete functioning of SMPS by having a look at the following figure.



Functional block diagram of SMPS

The SMPS is mostly used where switching of voltages is not at all a problem and where efficiency of the system really matters. There are few points which are to be noted regarding SMPS. They are

SMPS circuit is operated by switching and hence the voltages vary continuously.

The switching device is operated in saturation or cut off mode.

The output voltage is controlled by the switching time of the feedback circuitry.

Switching time is adjusted by adjusting the duty cycle.

The efficiency of SMPS is high because, instead of dissipating excess power as heat, it continuously switches its input to control the output.

Disadvantages

- There are few disadvantages in SMPS, such as
- The noise is present due to high frequency switching.
- The circuit is complex.
- It produces electromagnetic interference.

Advantages

- The advantages of SMPS include,
- The efficiency is as high as 80 to 90%
- Less heat generation; less power wastage.
- Reduced harmonic feedback into the supply mains.
- The device is compact and small in size.
- The manufacturing cost is reduced.
- Provision for providing the required number of voltages.

Applications

There are many applications of SMPS. They are used in the motherboard of computers, mobile phone chargers, HVDC measurements, battery chargers, central power distribution, motor vehicles, consumer electronics, laptops, security systems, space stations, etc.

Types of SMPS

SMPS is the Switched Mode Power Supply circuit which is designed for obtaining the regulated DC output voltage from an unregulated DC or AC voltage. There are four main types of SMPS such as

DC to DC Converter

AC to DC Converter

Fly back Converter

Forward Converter

5. PLC AND ITS APPLICATIONS

INTRODUCTION OF PROGRAMMABLE LOGIC CONTROLLER (PLC)

The Need for PLCs

PLC is used in the fully automated industries or plants or process, the actual processes handled and controlled by the controllers which are nothing but the programming logic controllers that means PLC plays a very important role in automation section.

PLCs constantly monitor the state of the systems through input devices and generate the control actions according to the logic given in the user program.

It is a heart of control systems, PLC monitors the state of the system through field input devices, feedback signals and based on the feedback signal PLC determine the type of action to be carried out at field output devices.

PLC provides easy and economic solution for many automation tasks like

- Operates control and monitoring.
- Co-ordination and communication.
- PID computing and control.
- Logic / sequence control.

Programmable Logic Controller

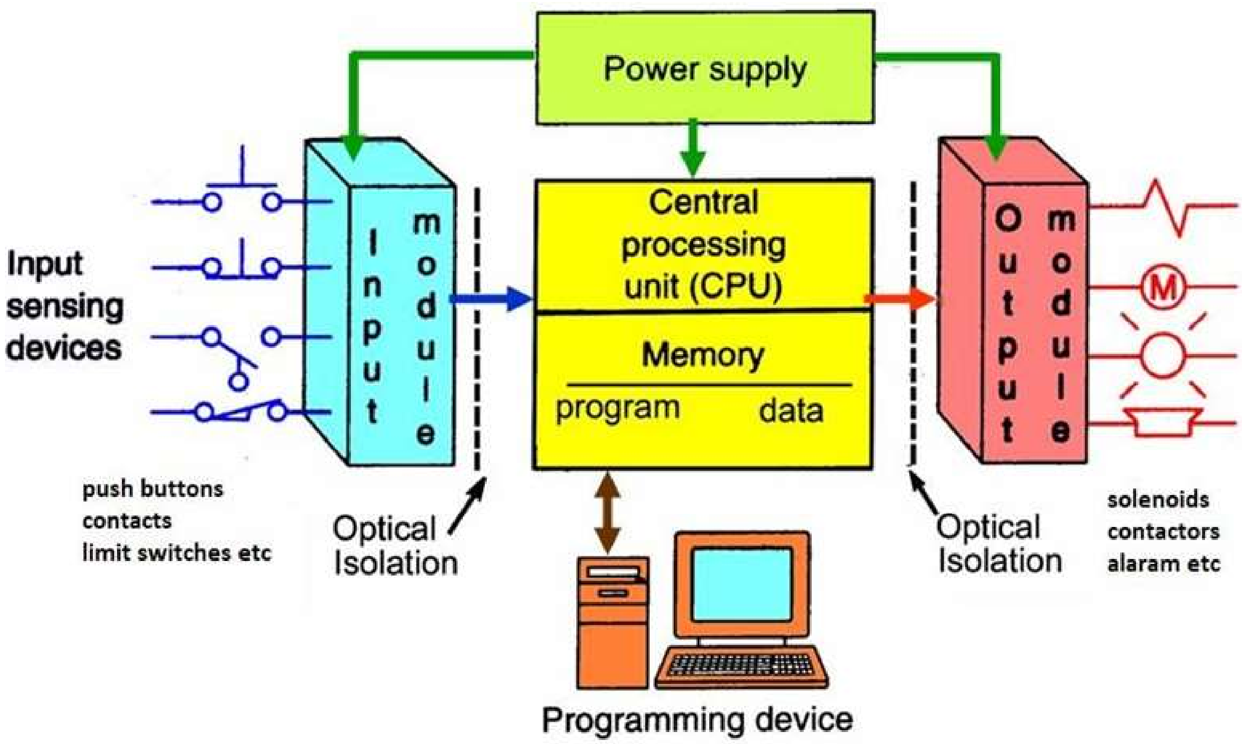
A programmable logic controller (PLC) is a specialized computer used to control machines and process.

It uses a programmable memory to store instructions and specific functions that include On/Off control, timing, counting, sequencing, arithmetic, and data handling

ADVANTAGES OF PLC

- Flexible
- Faster response time
- Less and simpler wiring
- Solid-state - no moving parts
- Modular design - easy to repair and expand
- Handles much more complicated systems
- Sophisticated instruction sets available
- Allows for diagnostics “easy to troubleshoot”
- Less expensive

DIFFERENT PARTS OF PLC BY DRAWING THE BLOCK DIAGRAM AND PURPOSE OF EACH PART OF PLC.



The block diagram of programming logic controller (PLC) is shown in above figure. The PLC has following basic sections are,

Processor section (CPU)

The processor section is brain of PLC which consists of RAM, ROM, logic solver and user memory. The central processing unit is heart of PLC. CPU controls monitors and supervises all operation within PLC. The CPU makes decision and executes control instructions based on the program instruction in memory.

Input and output module

The input module is a mediator between input devices and central processing unit (CPU) which is used to convert analog signal into digital signal.

The output module is a mediator between output devices and central processing unit (CPU) which is convert digital signal into analog signal.

Power supply

Power supply is provided to the processor unit, input and output module unit. Power supply may be integral or separately mounted unit. Most of the PLC operates on 0 volts DC and 24 volts.

Memory section

The memory section is the area of the CPU in which data and information is stored and retrieved. Data Memory is used to store numerical data required in math calculation, bar code data etc. User memory contains user's application program.

Programming device

Programming devices are dedicated devices used for loading the user program into the program memory or edit it and to monitor the execution of the program of the PLC. It is also used to troubleshoot the PLC ladder logic program. Hand held terminal (HHT) or dedicated terminal or personal computer are programming devices commonly used in most of the PLCs.

APPLICATIONS OF PLC

- In machining, packaging, material handling, automated assembly etc.
- In medical instruments and technologies
- In industrial manufacturing
- In Nano technology and robotics
- In automatic transfer switch
- In Traffic light Signal system
- It is used in civil applications such as washing machine, elevators working and traffic signals control.
- It is used to reducing the human control allocation of human sequence given to the technical equipments that is called **Automation**.
- It is used in batch process in chemical, cement, food and paper industries are sequential in nature, requiring time or event based decisions.

LADDER DIAGRAM

Ladder Diagram is a graphical programming language that you use to develop software for programmable logic controllers (PLCs). The most elementary objects in Ladder Diagram programming are contacts and coils, intended to mimic the contacts and coils of electromechanical relays.

Contacts and coils are discrete programming elements, dealing with Boolean (1 and 0; on and off; true and false) variable states.

Each contact in a Ladder Diagram PLC program represents the reading of a single bit in memory, while each coil represents the writing of a single bit in memory.

Discrete input signals to the PLC from real-world switches are read by a Ladder Diagram program by contacts referenced to those input channels.

DESCRIPTION OF CONTACTS AND COILS IN THE FOLLOWING STATES I) NORMALLY OPEN II) NORMALLY CLOSED III) ENERGIZED OUTPUT IV) LATCHED OUTPUT V) BRANCHING

Normally Open and Normally Closed electrical contacts make up electrical switches, relays, circuit breakers, and most any other electrical component that switches something on/off or can be switched on/off.

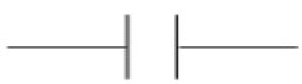
IMPORTANT CONCEPT:

Closed = Current flow

Open = No current flow

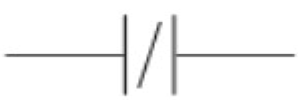
What is normally? This is simply the state that the contact is in when something else is not affecting it. If it is a relay then it is not energized. If it is a switch, then it is off. If it is a high limit such as a temperature alarm then the current temperature is below the limit.

NORMALLY OPEN - Is a contact that does not flow current in its normal state. Energizing it and switching it on will close the contact, causing it to allow current flow.



Normally Open
(NO)

Normally closed - Is a contact that flows current in its normal state. Energizing it and switching it on will open the contact, causing it to not allow current flow.



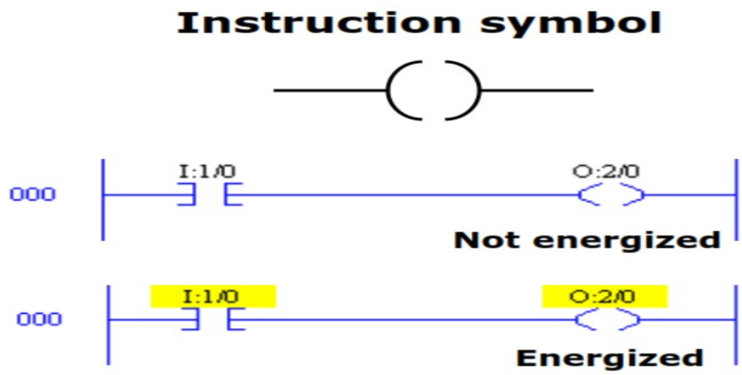
Normally Closed
(NC)

III) ENERGIZED OUTPUT

The OTE, also known as Output Energize, instruction will energize a single bit of data if the input leading to it is true. It's a fundamental instruction used in Programmable Logic Controllers (PLCs). This instruction will be found on the right side within a ladder logic structure and turn a bit to a HIGH state if the preceding instructions evaluate to true. If the same instructions evaluate to false, the OTE instruction will set the specified bit to a LOW state.

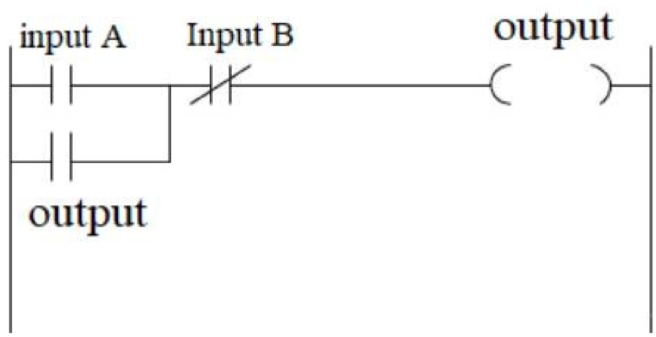
This Output energize (OTE) instruction is usually used in conjunction with XIC or XIO or any other input instruction in PLC.

If the logic preceding the OTE instruction is true (1), the OTE instruction will be energized.



IV) LATCHED OUTPUT

The latching is used where the output must be activated even after the entry ceases.



A simple example of such a situation is a motor, which is started by pressing a button switch. Although the switch contacts do not remain closed, it is required that the motor continue to run until a stop button switch is pressed. The latching used to stay the motor run until the push button is pressed again.

Circuits that are characteristic given the previous conditions are often needed in logic control. In this series the output is latched by using the output contact itself, so even though the input has changed, the output condition is fixed:

When there is an exit, another set of contacts associated with the exit is closed. These contacts form an OR logic gate system with the input contacts. Therefore, even if the A input is opened, the circuit will keep the output energized. The only way to release the output is to activate the normally closed contact B.

BRANCHING IN LADDER LOGIC

When two or more instructions are connected in parallel, it is called a branch. Ladder logic circuits almost always contain rungs with branches, and many have levels of branches. A branch level is assigned for each branch that is connected either in series or parallel.

1. SERIES BRANCH:

In the series branch, inputs or outputs are connected in the series.

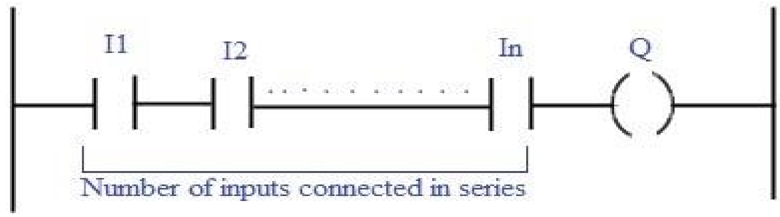


Figure: Representation of the Series Branch

2. PARALLEL BRANCH

In the parallel branch, inputs or outputs are connected parallelly.

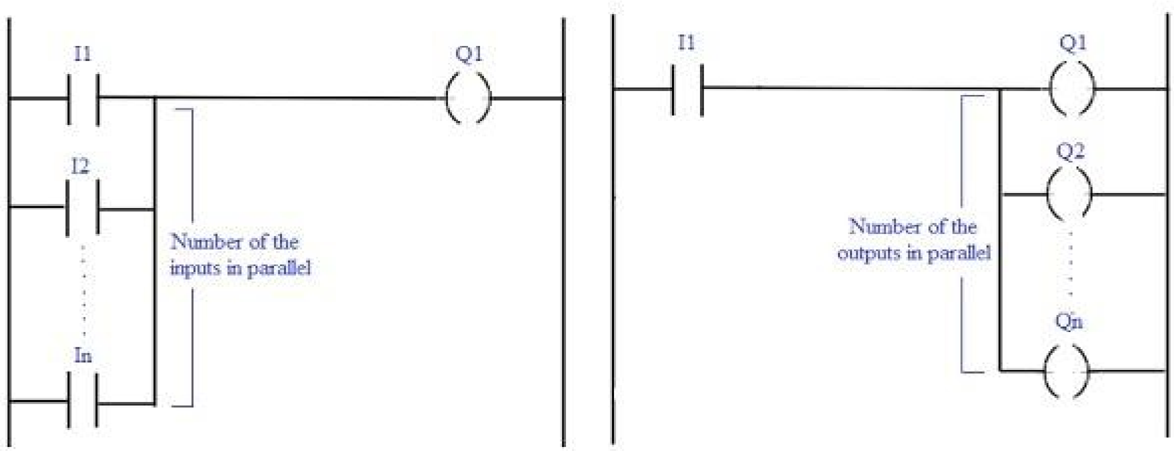


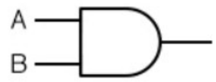
Figure: Representation of the Parallel Branch

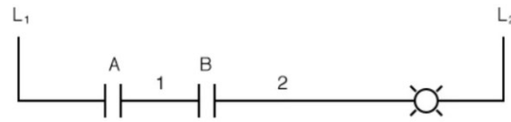
LADDER DIAGRAMS FOR I) AND GATE II) OR GATE AND III) NOT GATE.

We can mimic the AND logic function by wiring the two contacts in series Now, the lamp energizes only if contact A and contact B are simultaneously actuated.

A path exists for current from wire L1 to the lamp (wire 2) if and only if both switch contacts are closed.

A	B	Output
0	0	0
0	1	0
1	0	0
1	1	1



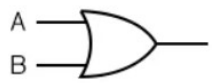


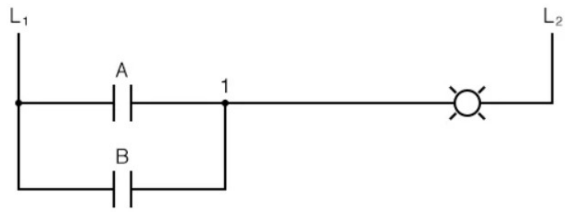
II) OR GATE

The lamp will come on if either contact A or contact B is actuated, because all it takes for the lamp to be energized is to have at least one path for current from wire L1 to wire 1.

What we have is a simple OR logic function, implemented with nothing more than contacts and a lamp.

A	B	Output
0	0	0
0	1	1
1	0	1
1	1	1

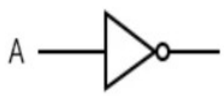


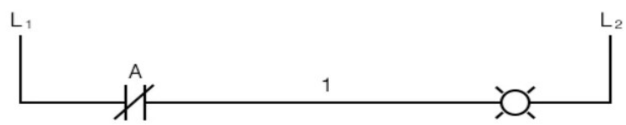


III) NOT GATE.

The logical inversion, or NOT, function can be performed on a contact input simply by using a normally-closed contact.

A	Output
0	1
1	0

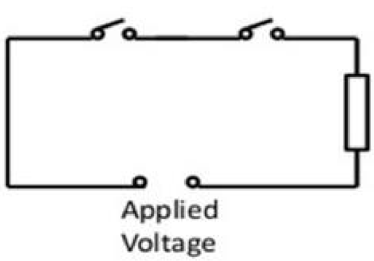




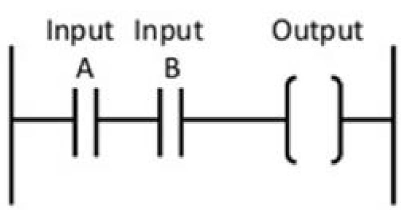
Now, the lamp energizes if the contact is not actuated, and de-energizes when the contact is actuated.

LADDER DIAGRAMS FOR COMBINATION CIRCUITS USING NAND, NOR, AND, OR AND NOT

AND Gate:

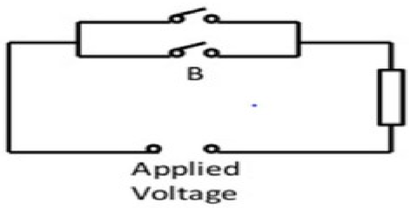


(a)

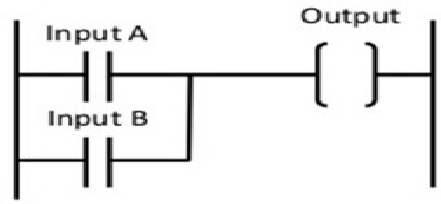


(b)

OR Gate:

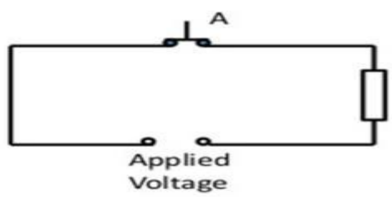


(a)

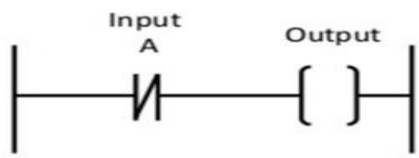


(b)

NOT Gate:

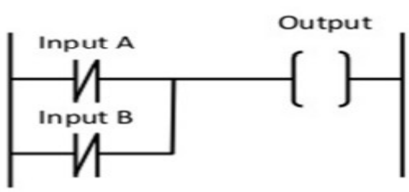


(a)

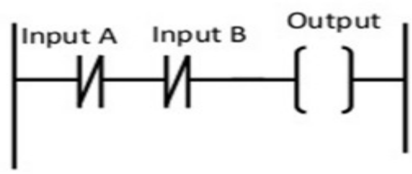


(b)

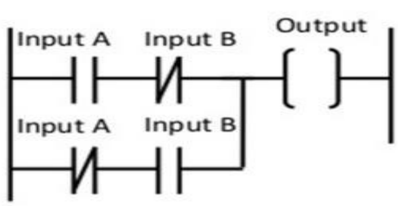
NAND Gate:



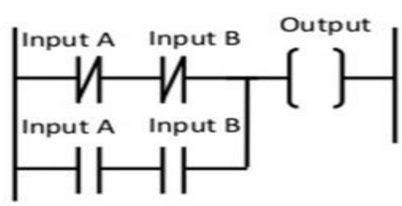
NOR Gate



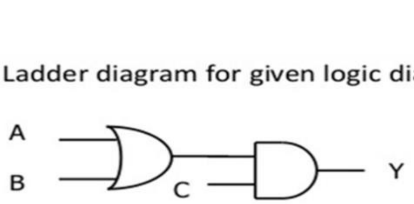
XOR Gate:

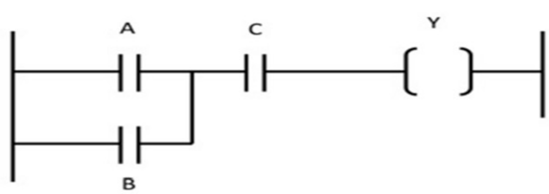


X-NOR Gate

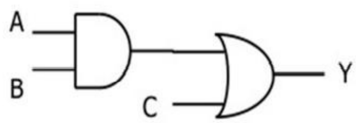


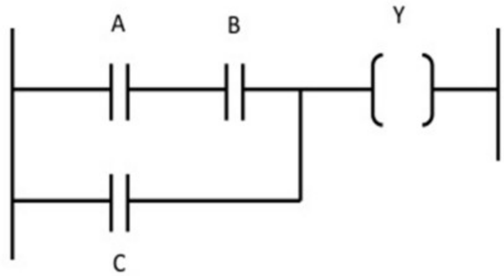
Draw Ladder diagram for given logic diagram

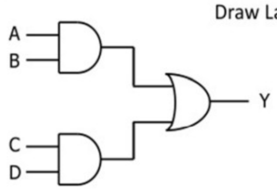




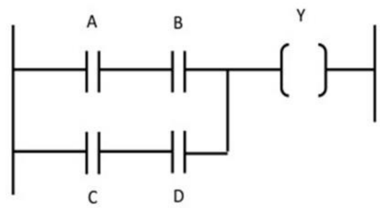
Draw Ladder diagram for given logic diagram



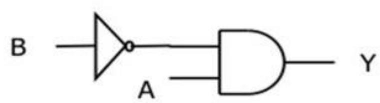




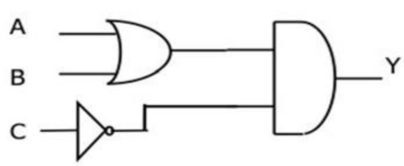
Draw Ladder diagram for given logic diagram

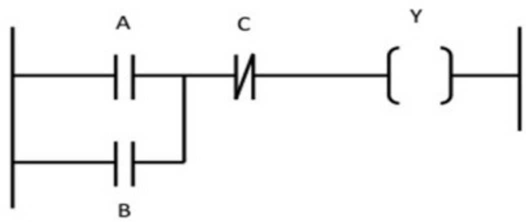


Draw Ladder diagram for given logic diagram



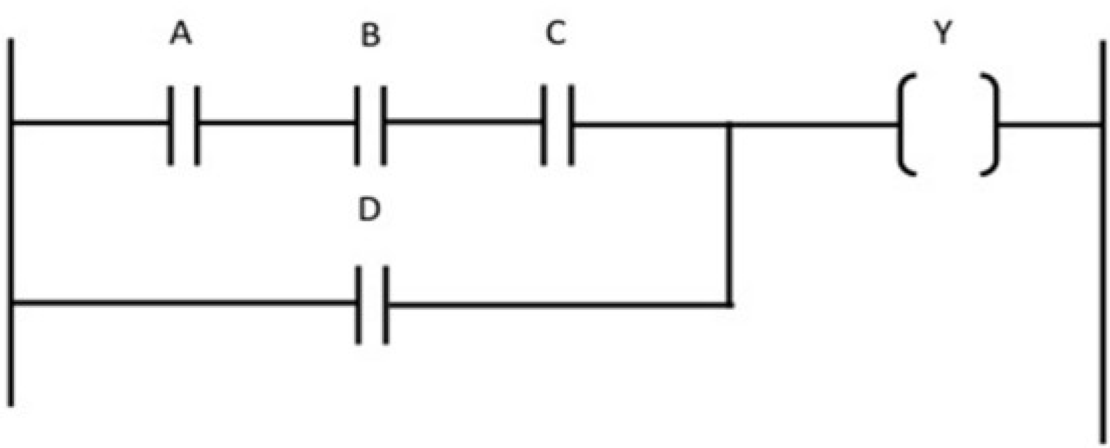
Draw Ladder diagram for given logic diagram





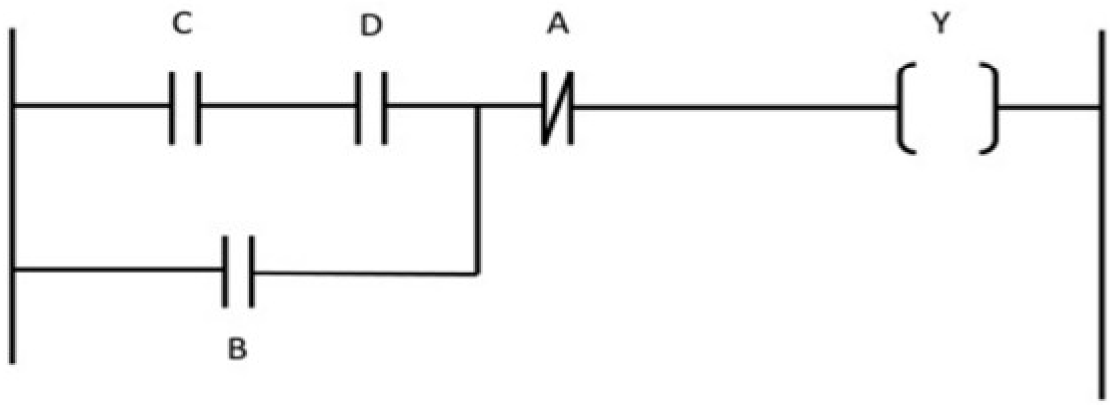
Draw Ladder diagram for given Boolean Expression

$$Y = ABC + D$$



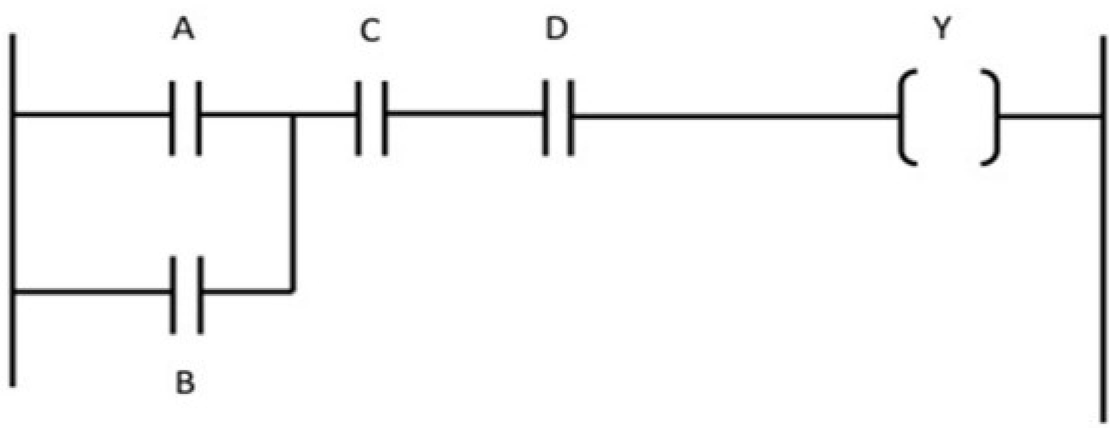
Draw Ladder diagram for given Boolean Expression

$$Y = \overline{A}(B + CD)$$



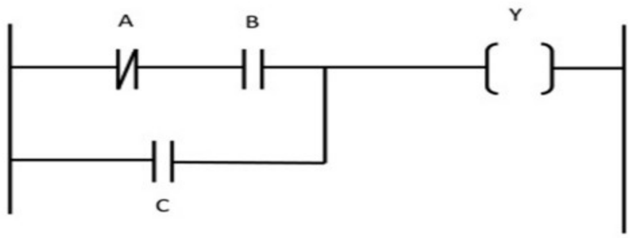
Draw Ladder diagram for given Boolean Expression

$$Y = (A + B)CD$$



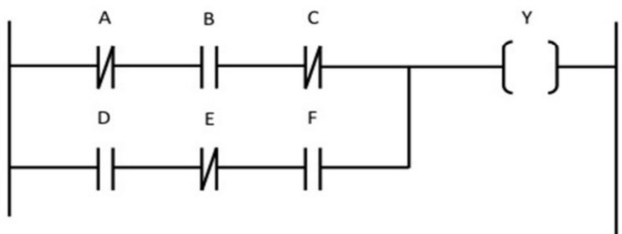
Draw Ladder diagram for given Boolean Expression

$$Y = \overline{A}B + C$$



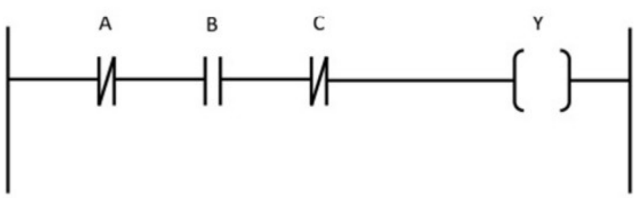
Draw Ladder diagram for given Boolean Expression

$$Y = (\overline{A}B\overline{C}) + (D\overline{E}F)$$



Draw Ladder diagram for given Boolean Expression

$$Y = (\overline{A} + B) + (\overline{A} + B + \overline{C})$$



TIMERS-I) T ON II) T OFF AND III) RETENTIVE TIMER

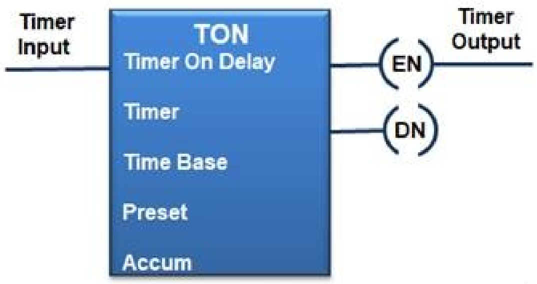
1. On Timer (TON)

What is TON?

TON is called 'On Delay Timer'.

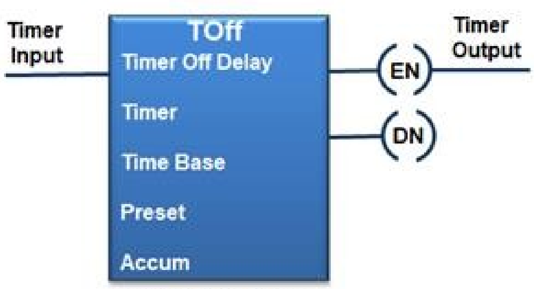
An on-delay timer (TON) is a programming instruction which use to start momentary pulses for a set period of time.

Let's see, a simple construction of the AB PLC On-delay timer programming instruction.



On Delay PLC Timer

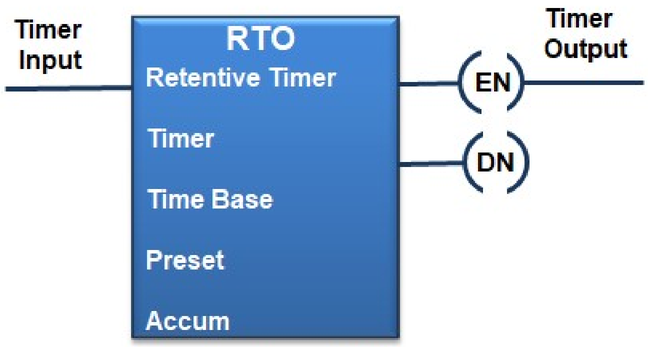
In the LD programming, when an On-delay timer is energized (True), it delays turning 'on' the timer's output.
This output will be 'on' until the timer's preset time value is reached.off delay PLC timer instruction. It helps to activate the output (like machine or process) contact based on the delay time.
Example: Running Electric Motor after 10 seconds.
– If you press the button (NC), it starts the momentary pulse. After the 10 seconds, the motor will be 'On'.
2. Off Timer (TOFF)
What is TOFF?
TOFF is also known as an 'Off-Delay Timer'.
A off-delay (TOF) timer is a PLC programming instruction which use to switch off the output or system after a certain amount of time.
See here, a basic structure of AB PLC Off delay timer programming instruction.



Off Delay PLC Timer

In PLC programming, when the off-delay timer is energized (True), it immediately turns 'on' its output. The output will be 'on' till it reaches the setting time.
When it reaches preset time, the output turns 'off'. Due to the turning 'off' condition, the timer is de-energized (False).
It helps to delay the shut down of machinery or process in automation industries.
Example: Stop electric motor after the 10 seconds.
– Firstly, you should switch press (NC), the motor will 'on' for the 10 seconds. After 10 seconds, the motor will automatically stop (NO).
This is the main function of the off-delay timer.

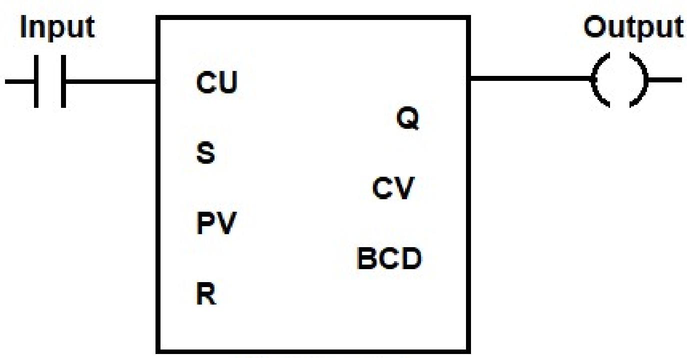
3. Retentive Timer On/Off (RTO)
The main function of the RTO is used to hold or store the set (accumulated) time.
RTO is used in the case when there is a change in the rung state, power loss, or any interruption in the system.
In the AB PLC, retentive timer instruction looks like this.



Retentive Timer On/Off

COUNTERS-CTU, CTD

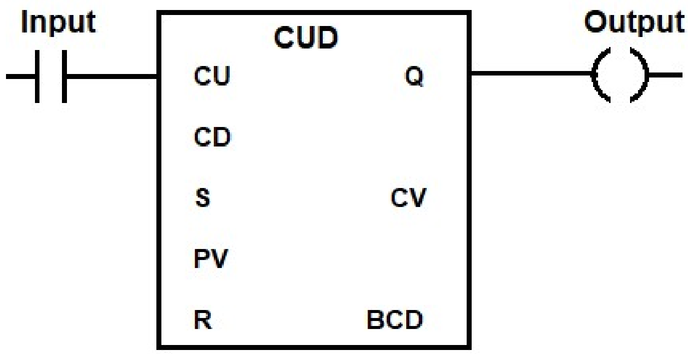
Up Counter
Up counter counts from zero to the preset value. Basically, it increases the pulse or number.
Up counter is known as the ‘**CTU**’ or ‘**CNT**’ or ‘**CC**’ or ‘**CTR**’.
Up counter function block diagram:



Up Counter

We can also set the initial and target value as an input to the counter.
Here, the up-counter in PLC can count the value from the initial value to the target value. This initial value must be less than the target value. Most of the time, it is set as zero.

Down Counter
The down counter counts from the preset value to zero. It decreases the pulse or number.
Down counter is shortly known as the ‘**CTD**’ or ‘**CD**’.
Down counter function block diagram:



Up-Down Counter

The down counter counts from target value to the initial value by decreasing it. This initial value must be less than the target value.

Counter Instruction Addressing for Siemens PLC

In Siemens PLC, up, down and up-down counters are used. These three PLC counters require some important factors

- S – Set the value of a counter.
- Q – Output of the counter.
- R- Reset value of the counter.
- PV- Preset counter value.
- CV – Count Variable.
- BCD – Current count in binary decimal code.
- Preset counter value (PV) and Count Variable (CV) require the same addressing format. The standard addressing format of the PV and CV in LD.

Summary of PLC Counter Function

- The basic counter function is to count the digital signal pulse or binary system.
- Different PLC brands offer a different range of counter values.
- Counters work as per the supported mode.
- Counter operates in up mode, down mode, bidirectional mode, and quadrature mode.
- Up counting starts from the zero or initial value to the target value.
- Down counting starts from target value down to the initial value.

LADDER DIAGRAMS USING TIMERS AND COUNTERS

LADDER DIAGRAMS USING TIMERS

In many of the PLC control tasks, there is a need to control the time an example of this will be using a PLC to control a motor. The motor would require to be controlled to operate for a certain interval, and that’s why PLCs have timers and the timers are built-in devices in a PLC. By using he internal CPU clock the timer would count the time. Different PLC timers are programmed in different ways, so we can consider a timer to act as a relay with coils, which would open or close when it is energized according to the pre-set time.

Timer instructions

The timer instructions are the output instructions which is used to time the intervals for which the rung conditions are true or false. The timer accuracy will be depended upon the microprocessor which is being used. The timer instruction is composed of two values and they are

- **Accumulated value** – This is a current number of time-based intervals that have been counted from the moment when the timer is energized.
- **Preset value** – This value is set by the programmer, if the preset value is less than or equal to the accumulated value then a status bit is set and this bit is to control an output device.
- Each timer is composed of two status bit
- **Timer enable-bit** – This bit will be set if the rung condition to the left of the timer instruction is true and when this bit is set then the accumulated value will be incremented on each time base interval till it reaches the preset value.
- **Done bit** – This bit will be set if the preset value and the accumulated value are equal and it will be reset if the rung condition is false.

Timer working

The timer will be activated if the execution condition is started and it will be reset if the execution condition stops or goes OFF. If the execution condition keeps ongoing or if it is long enough for the timer to time down to zero. Then the completion flag will be turned ON and it will remain ON till the execution condition is completed or turn off.

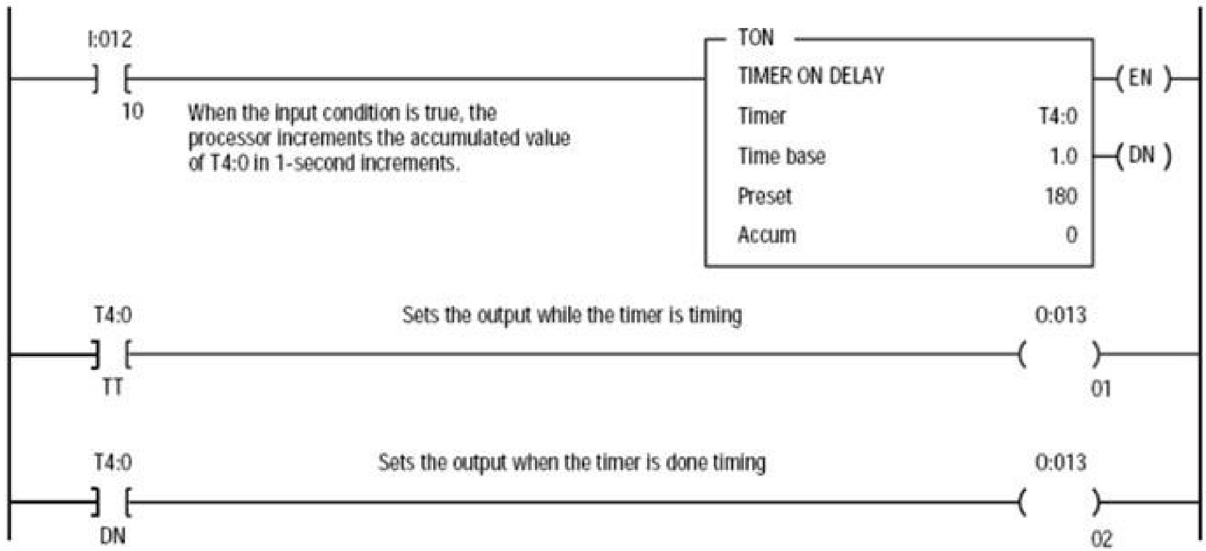
The address of the timer is unique in the PLC memory, the timer instruction is one element and the timer element is composed of 16-bits. The word zero will cover the status bits and it has three state bits such as EN, TT, and DN. The word one is for the preset value and the word two is for the accumulated value.

Types of timers

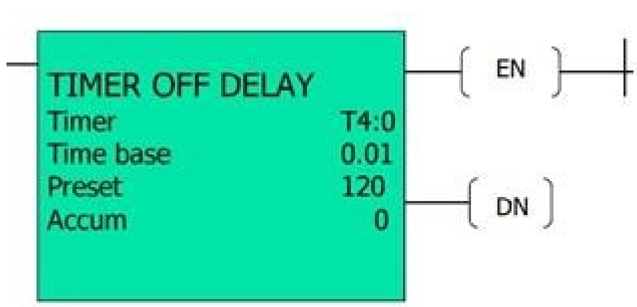
We can see different types of timers in a PLC

ON delay timer

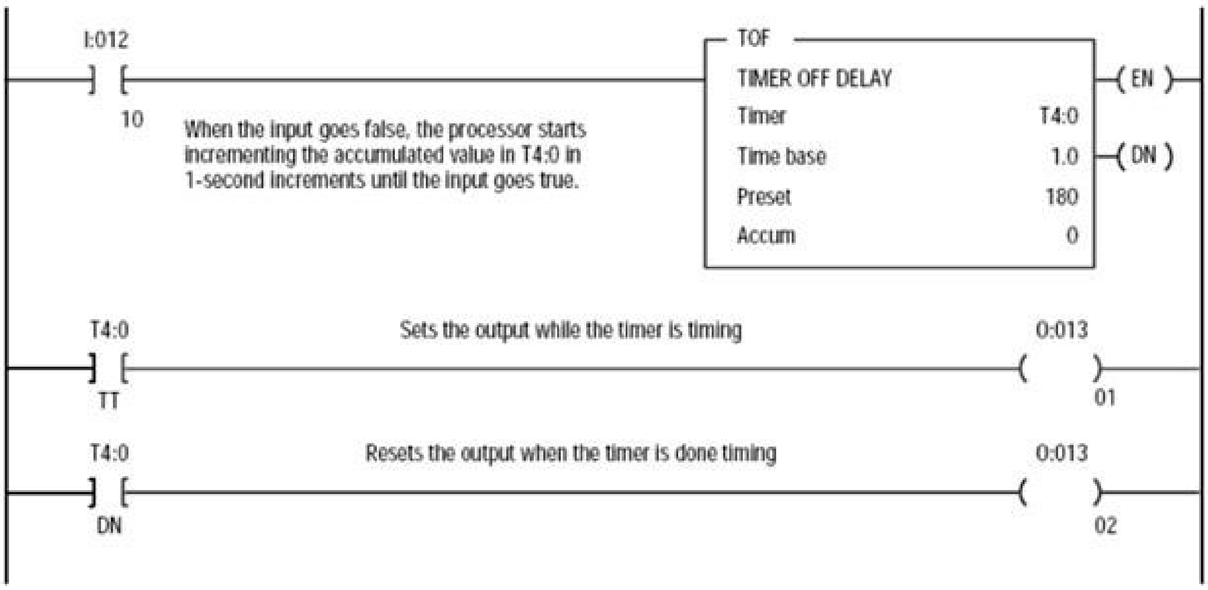
This type of timer can be seen in small PLC's and this timer would be on after a particular time delay. This is widely used for PLC programming we can also create other timer functions by using an on-delay timer. So this timer would delay the turning ON time in a system. This type of timer would count the time base interval if the instruction is true. So if the rung condition becomes true then the on delay timer instructions will start to count the time base intervals. So if the rung conditions stay true then the timer would adjust its accumulated value during each evaluation till it reaches the preset value. So when the rung condition becomes false then the accumulated value is reset. The three timer bits EN, TT, and DN can be used as rung conditions.



OFF delay timer



This is the exact opposite of the on delay timer; this timer would delays the turning off. So these timers are on for a fixed period of time before turning off. The output will be turned off after a delay, so when this timer is turned on then the output is also turned on. So if the output needs to be turned off then it needs to be turned on in the beginning. So this timer won't be activated before we turn the input off again, so if we do that the timer will start the count after the delay and the output will be turned off.



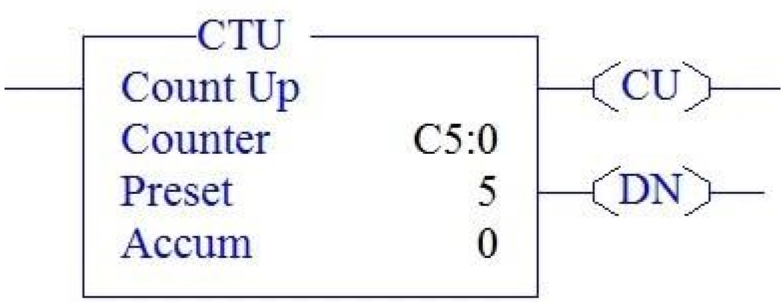
Pulse timer

This timer is not used widely in a PLC it can be very helpful for some PLC operations. The pulse timer is used to generate the pulse and it will be of a specific length. The pulse timer can be activated by turning on the input and when it is turned on then the timer would start counting the time. So basically a pulse timer can switch on or off for a fixed period of time.

LADDER DIAGRAMS USING COUNTERS

To study the working of **Up Counter PLC program** in Allen Bradley Programmable Logic Controller (PLC).

Up Counter

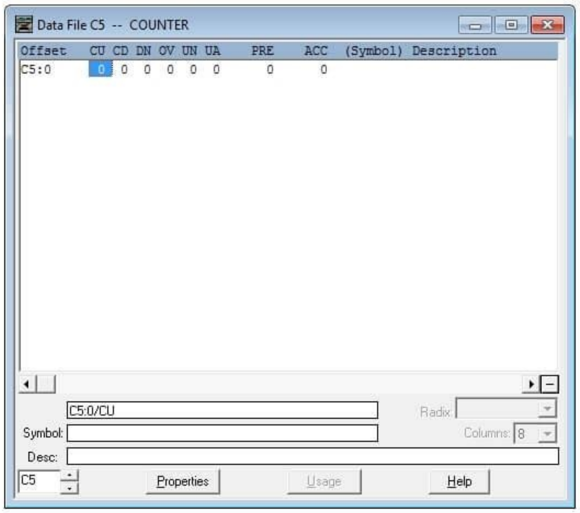


In the above picture, there are totally three parameter,

COUNTER: C4:0 – Counter File name (Timer C5:0, C5:1, C5:2…)

PRESET –PRE: Limit value of COUNT-Up to how much it should count

ACCUMULATOR –ACC: Running Value of counter when condition turn ON.

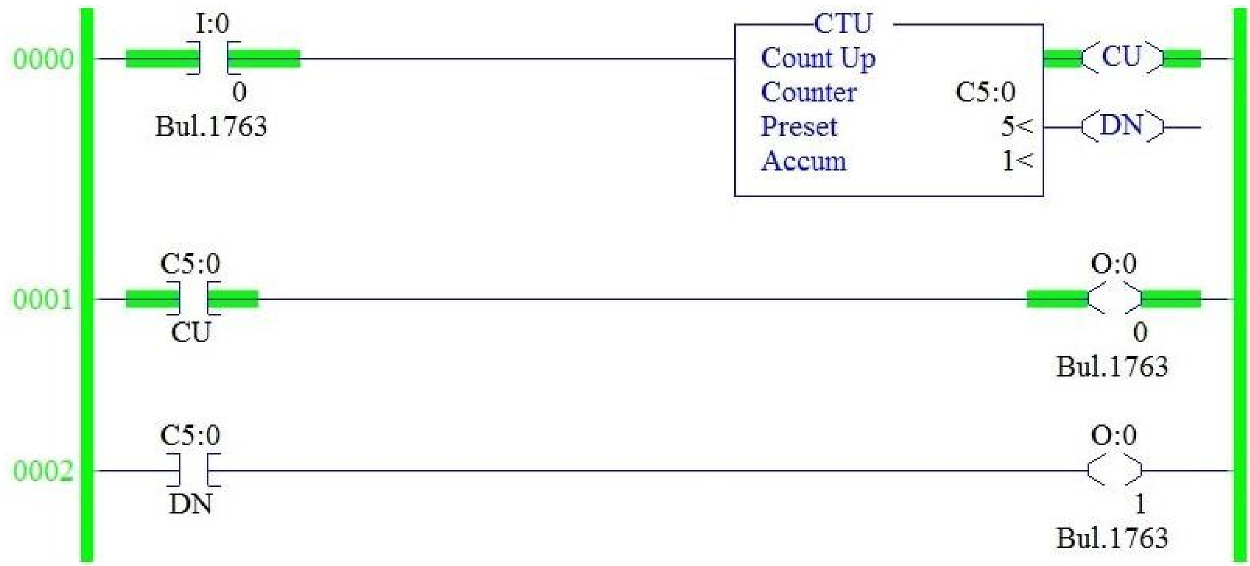


From the data file, along with preset and accumulator, we have few more bits,
CU: Count up Bit-whenever the counter is enable makes this bit to go ON.
DN: Done Bit-When accumulator value reached preset value, done bit turns to ON.
OV: Over Flow Bit-When accumulator value reached the limit value (32767),it rolls back to -32767 for the upcoming counter operation, Overflow bit turns ON, in this condition.

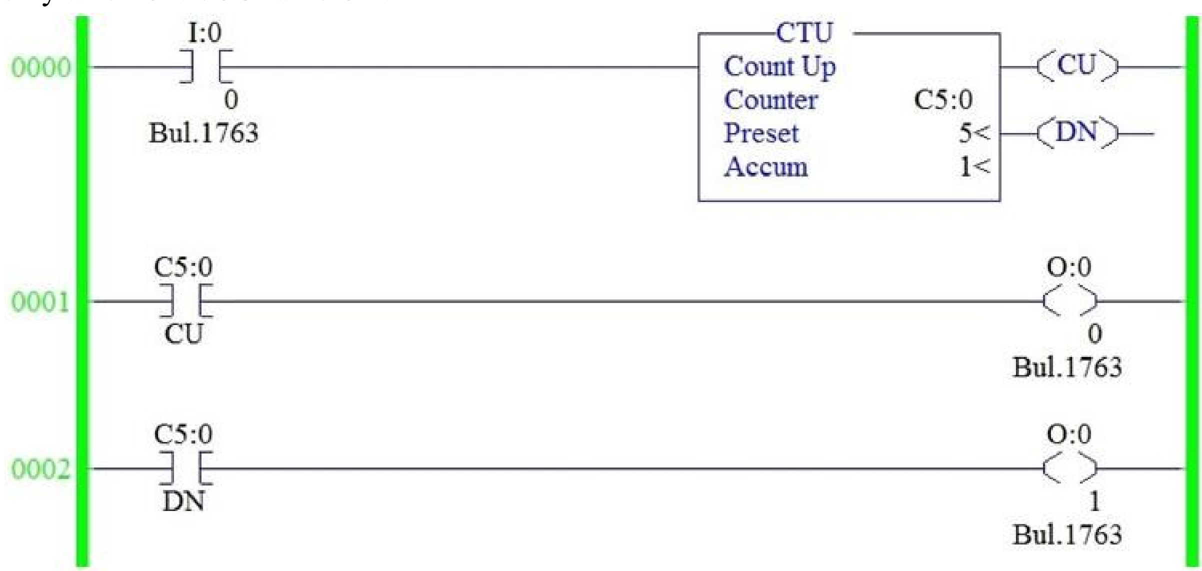
Notes:
UA-Update Accumulator Value-Only used when high speed counters are used in the program.
CD & UN-Used for down Counter Function.

Up Counter Description Using PLC Program

I:0/0 is used to give input to counter and Preset value is set to 5.
Counter Count up Bit (C5:0/CU)
In the below Ladder logic,
Rung 000 – Having condition input I:0/0 which gives input to counter to perform counter function.
Rung 0001 – Having Counter CU Bit which enable only when counter is in function or when input to the counter turns ON.

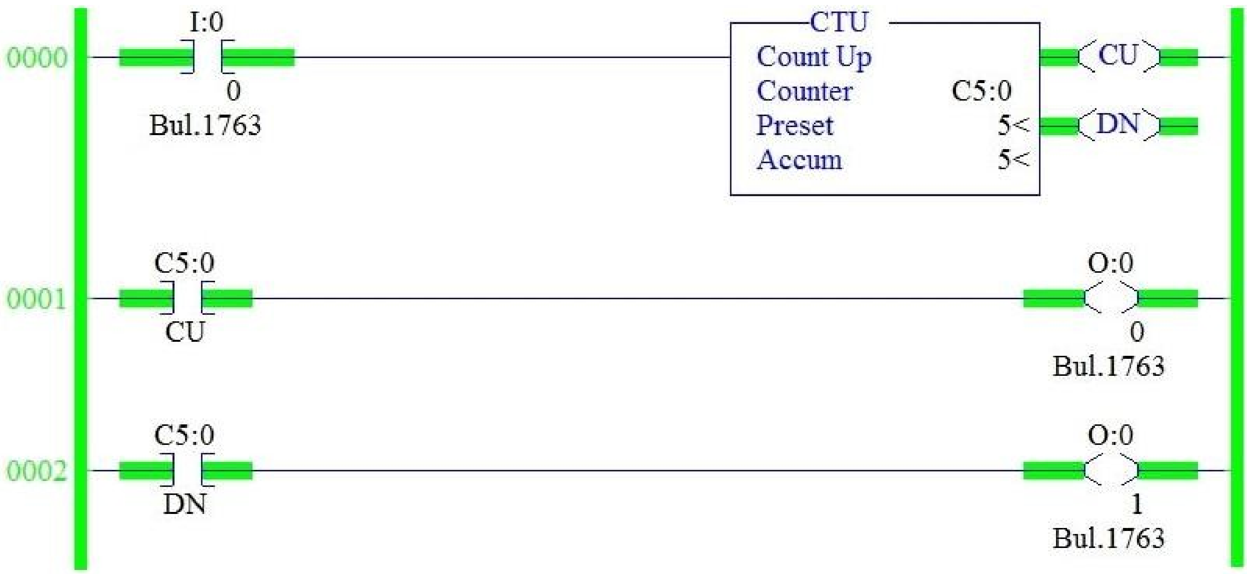


In the below ladder Logic,
When input to the counter turn OFF (I:0/0),Counter CU bit turns OFF. Output O:0/0 turns ON only when C5:0/CU turns ON.



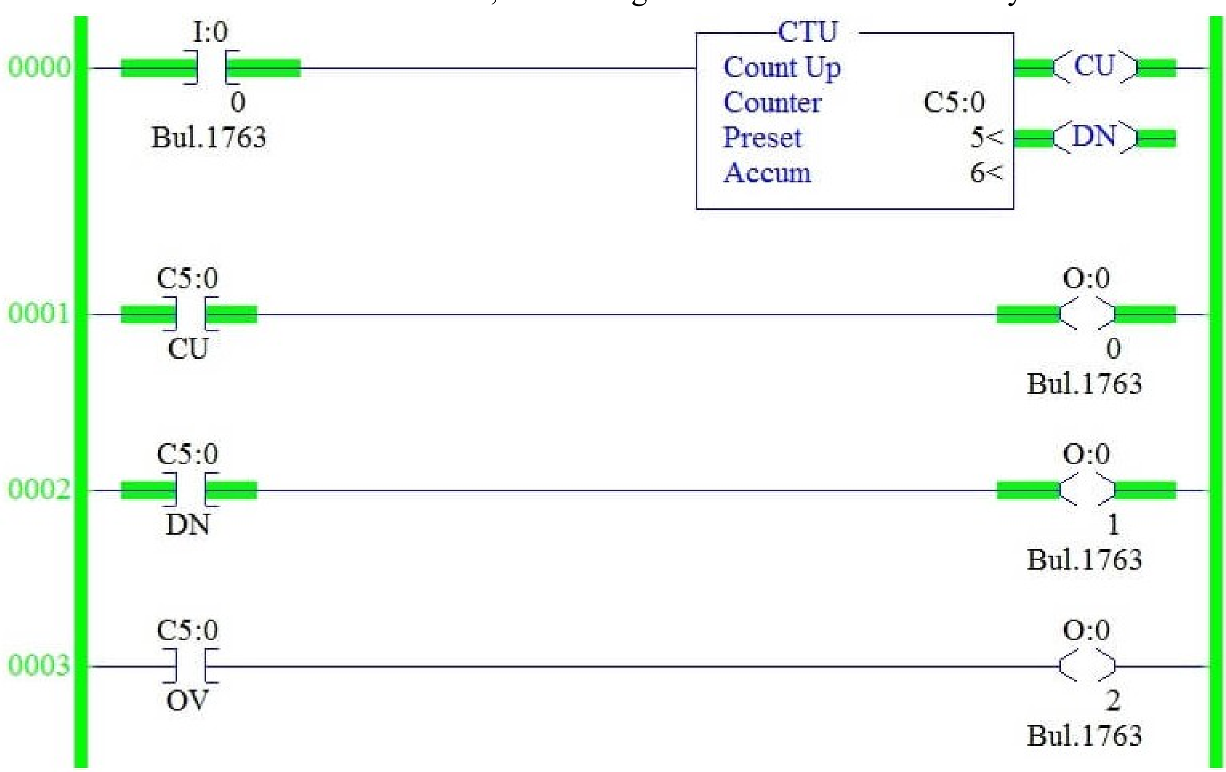
Counter Done Bit (C5:0/DN)

In the Below Ladder Logic,
When accumulator value reaches the Preset, Counter Done bit (Cu5:0/DN) turns ON.



In the below ladder Logic,

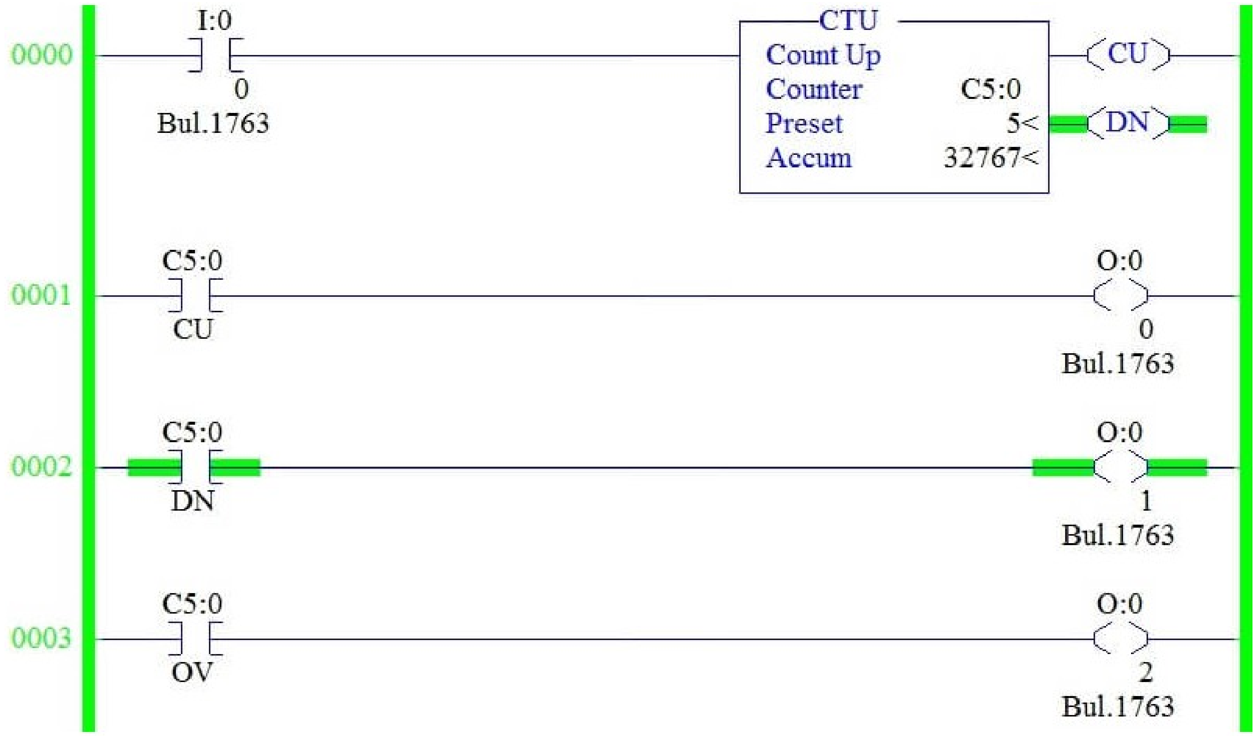
Done bits remains in the ON condition, even though accumulator value runs beyond Preset.



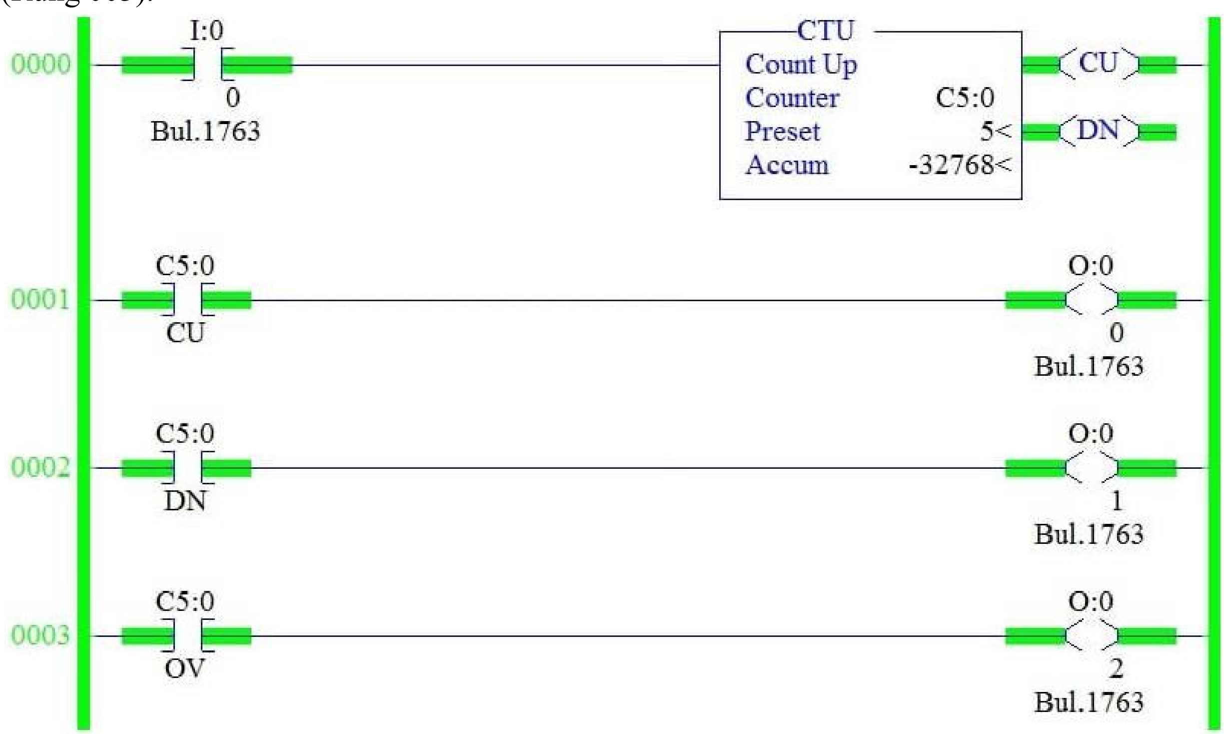
Counter Overflow Bit (C5:0/OV)

In the Below Ladder Logic,

Counter accumulator value overflows when accumulator value reaches 32767 in Allen Bradley PLC Programming.



In the below ladder Logic,
When we turn ON the I:0/0 for the 32768 time, accumulator value rolls back to -32768 and start counting from -32767 to 32767.Counter Overflow bit turns ON when this condition happen (Rung 003).



Conclusion:

We can use this explanation to understand the working of Up Counter function in Allen Bradley Programmable Logic Controller (PLC).

PLC INSTRUCTION SET

Naming Convention

During the development of a PLC program, we must use specific names to identify the inputs, outputs, memory flags, timers, and counters.

PLC manufacturers use a variety of approaches in naming the inputs, outputs and other resources.

A typical naming convention is to identify inputs with the letter “I” and outputs with the letter “O”, followed by a 1-digit number that identifies the slot number and a 2-digit number that identifies the position of the input or output in the slot.

For example:

- I1:00 refers to the first input of slot 1
- O2:00 refers to the first output of slot 2.

Naming Convention

During the development of a PLC program, we must use specific names to identify the inputs, outputs, memory flags, timers, and counters.

PLC manufacturers use a variety of approaches in naming the inputs, outputs and other resources.

A typical naming convention is to identify inputs with the letter “I” and outputs with the letter “O”, followed by a 1-digit number that identifies the slot number and a 2-digit number that identifies the position of the input or output in the slot.

For example:

- I1:00 refers to the first input of slot 1
- O2:00 refers to the first output of slot 2.

Examine if Closed (XIC)

If the input device is ON or closed, then the corresponding bit in the data memory (input image) is set to true, thus allowing (conceptually) the energy to flow from its left side to its right-hand side.

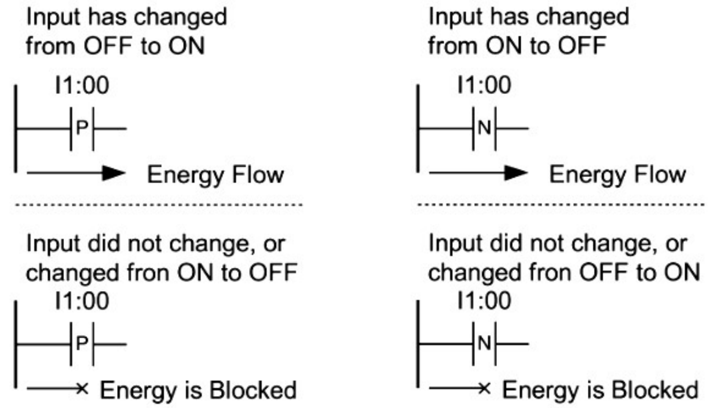
Otherwise, it is set to false, thus blocking the energy.

Examine if Open (XIO)

If the input device is OFF or Open, then the corresponding bit in the data memory (input image) is set to true, thus allowing (conceptually) the energy to flow from its left side to its right-hand side.

Otherwise, it is set to false, thus blocking the energy.

Input Transition Sensing Instructions



Positive Transition Sense (PTS) Negative Transition Sense (NTS)

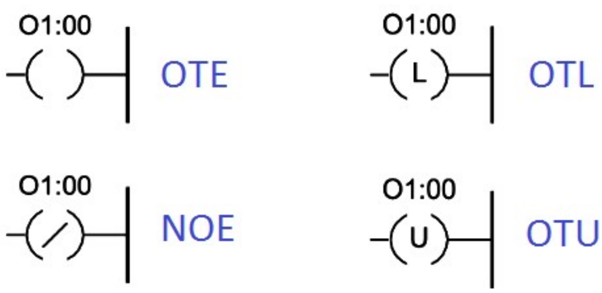
Positive Transition Sense (PTS)

The condition of the right link is ON for one ladder rung evaluation when a change from OFF to ON at the specified input is sensed.

Negative Transition Sense (NTS)

The condition of the right link is ON for one ladder rung evaluation when a change from ON to OFF at the specified input is sensed.

Output Instructions



Output Energize (OTE)

If the condition of the left link of the OTE is ON then the corresponding bit in the output data memory is set. The device wired to this output is also energized.

Negative Output Energize (NOE)

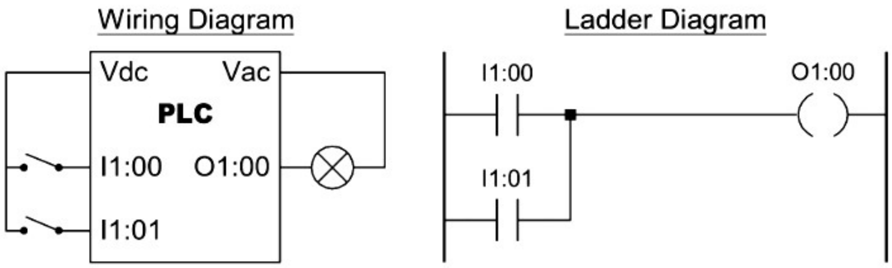
If the condition of the left link of the OTE is OFF then the corresponding bit in the output data memory is set. The device wired to this output is also energized.

Output Latch/Set and Output Unlatch/Reset (OTL), (OTU)

If the condition of the left link of the OTL is momentary ON then the corresponding bit in the output data memory is set, and remains set even if the condition switches to the OFF state. The output will remain set until the condition of the left link of the OTU is momentary ON.

Basic Logic Functions

Two Input OR Function

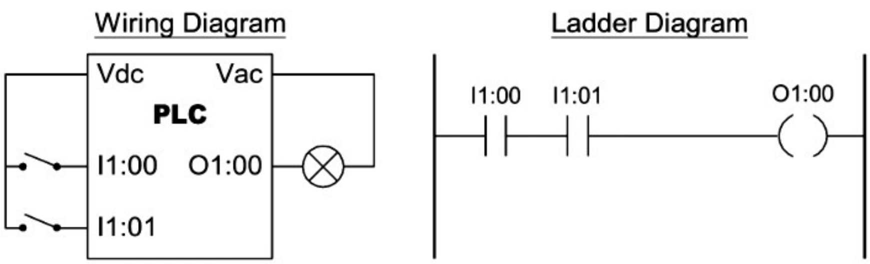


Truth Table

I1:00	I1:01	O1:00
OFF	OFF	OFF
OFF	ON	ON
ON	OFF	ON
ON	ON	ON

The output is ON if any of the two inputs is ON.

Two Input AND Function

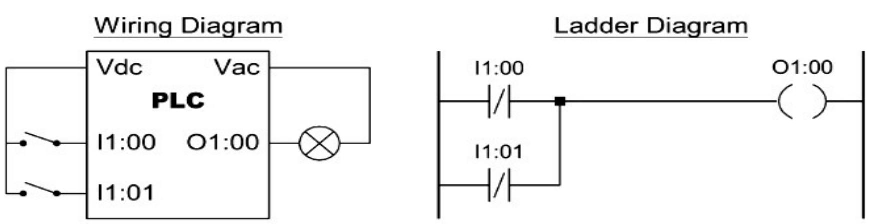


Truth Table

I1:00	I1:01	O1:00
OFF	OFF	OFF
OFF	ON	OFF
ON	OFF	OFF
ON	ON	ON

The output is ON if both of the two inputs are ON.

Two Input NAND Function

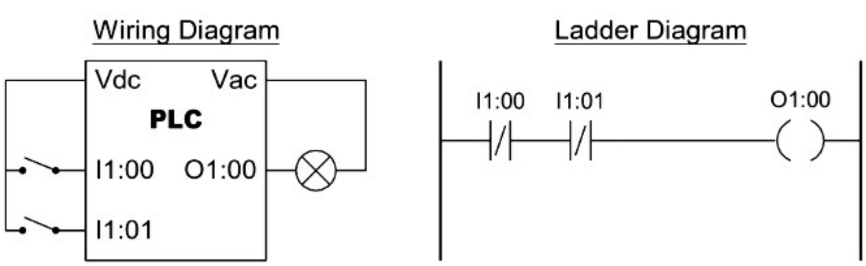


Truth Table

I1:00	I1:01	O1:00
OFF	OFF	ON
OFF	ON	ON
ON	OFF	ON
ON	ON	OFF

The output is ON if any of the two inputs is OFF.

Two Input NOR Function

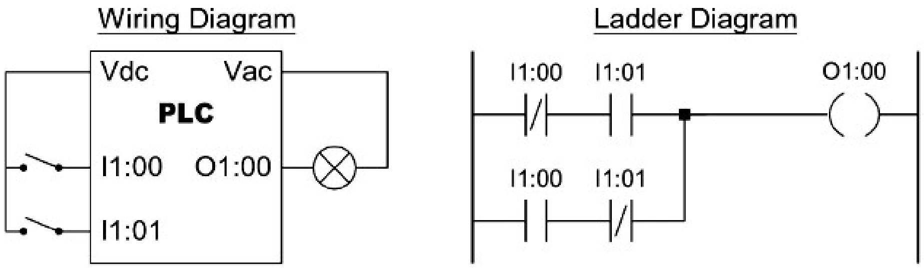


Truth Table

I1:00	I1:01	O1:00
OFF	OFF	ON
OFF	ON	OFF
ON	OFF	OFF
ON	ON	OFF

The output is ON if both of the two inputs are OFF.

Two Input EXOR Function

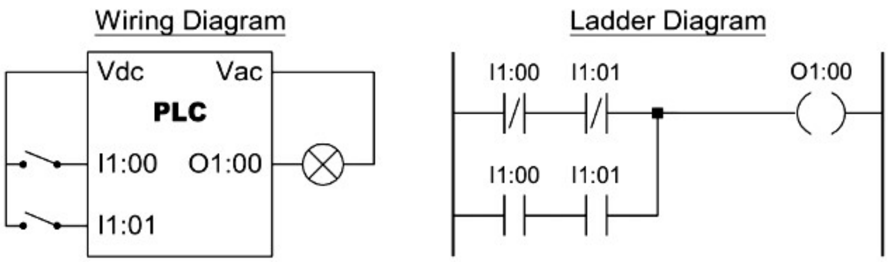


Truth Table

I1:00	I1:01	O1:00
OFF	OFF	OFF
OFF	ON	ON
ON	OFF	ON
ON	ON	OFF

The output is ON if any of the two inputs is ON, but not both.

Two Input EXNOR Function



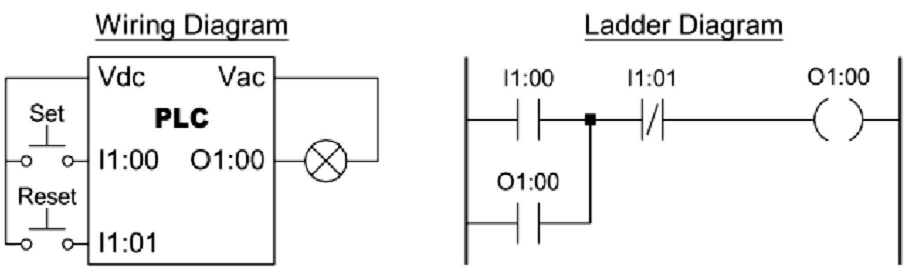
Truth Table

I1:00	I1:01	O1:00
OFF	OFF	ON
OFF	ON	OFF
ON	OFF	OFF
ON	ON	ON

The output is ON if both of the two inputs are either OFF or ON.

Set/Reset Latch Instructions

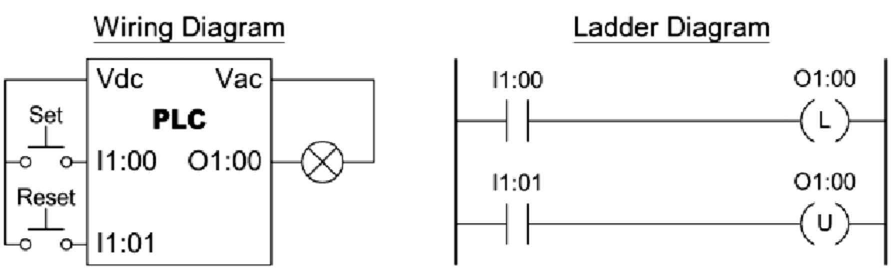
Set/Reset Latch using a Hold-in contact



Truth Table

I1:00	I1:01	O1:00
OFF	OFF	O1:00'
OFF	ON	OFF
ON	OFF	ON
ON	ON	OFF-?

Set/Reset Latch using Latch/Unlatch outputs



Truth Table

I1:00	I1:01	O1:00
OFF	OFF	O1:00'
OFF	ON	OFF
ON	OFF	ON
ON	ON	OFF-?

Notes:

- O1:00’ means that the output is unchanged
- If both inputs are ON then normally the output is OFF, since the Unlatch rung appears last in the ladder diagram.

Timer Instructions

Timer Instructions are output instructions used to time intervals for which their rung conditions are true (TON), or false (TOF).

These are software timers. Their resolution and accuracy depend on a tick timer maintained by the microprocessor.

Each timer instruction has two values (integers) associated with it:

- Accumulated Value (ACC): This is the current number of ticks (time-base intervals) that have been counted from the moment that the timer has been energized.
- Preset Value (PR): This is a predetermined value set by the programmer. When the accumulated value is equal to, or greater than the preset value, a status bit is set. This bit can be used to control an output device.

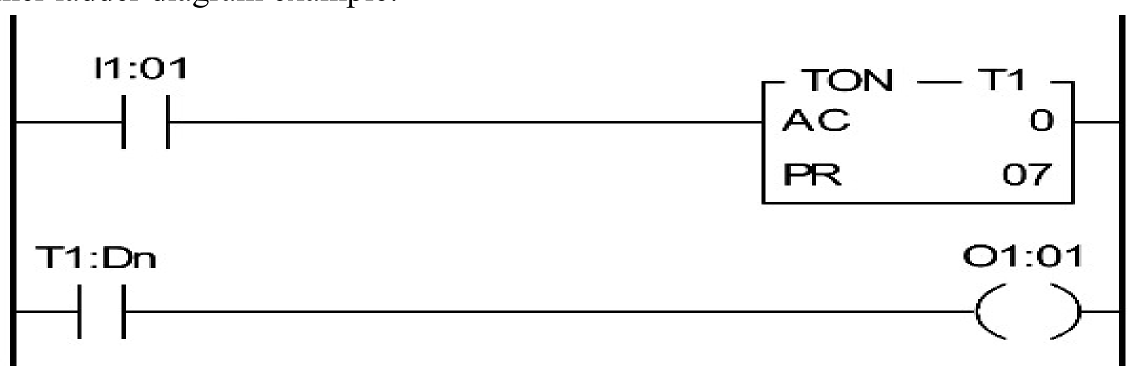
Each timer is associated with two status bits:

- Timer Enable Bit (EN): This bit is set when the rung condition to the left of the timer instruction is true. When this bit is set, the accumulated value is incremented on each time-base interval, until it reaches the preset value.
- Done Bit (DN): This bit is set when the accumulated value is equal to the preset value. It is reset when the rung condition becomes false.

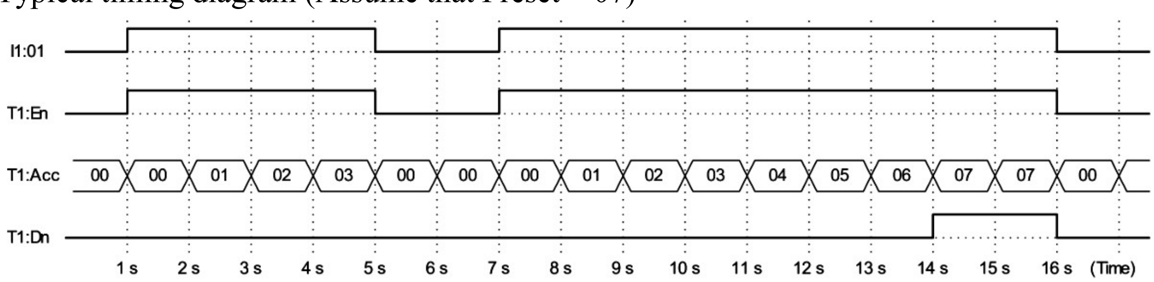
Timer On-Delay (TON) Instruction

The TON instruction begins to count when its input rung conditions are true. The accumulated value is reset when the input rung conditions become false.

Timer ladder diagram example:



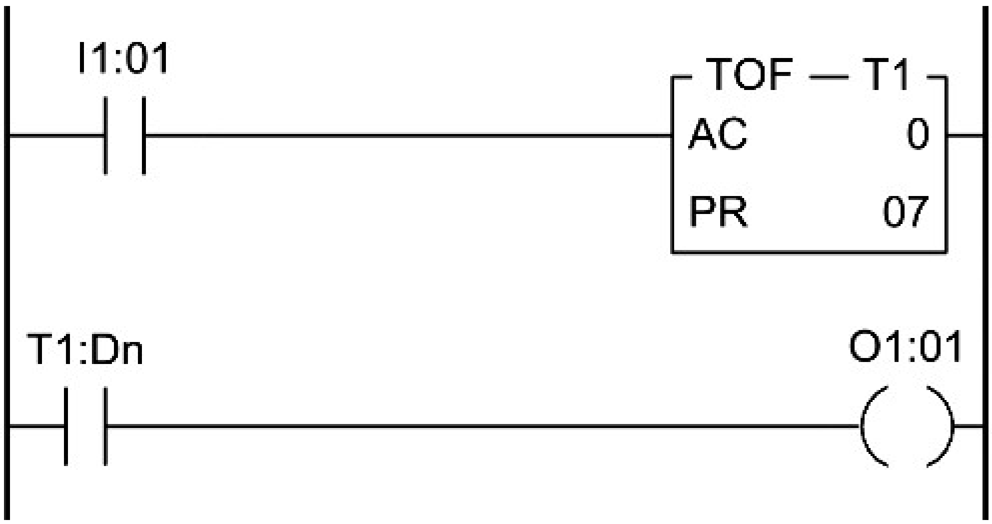
Typical timing diagram (Assume that Preset = 07)



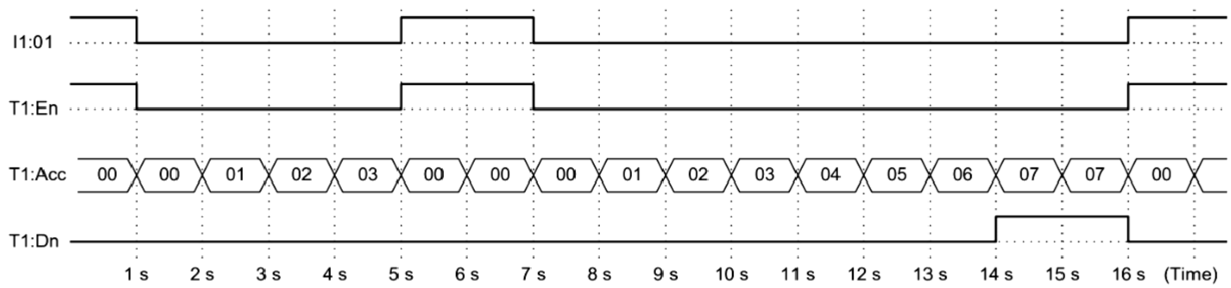
Timer Off-Delay (TOF) Instruction

The TOF instruction begins to count when its input rung makes a true-to-false transition, and continues counting for as long as the input rung remains false. The accumulated value is reset when the input rung conditions become false.

Timer ladder diagram example:



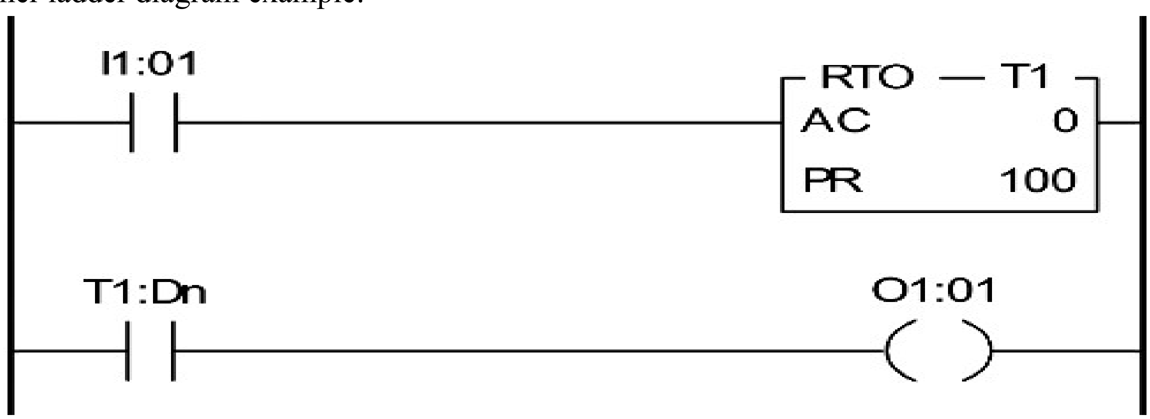
Typical timing diagram (Assume that Preset = 07)



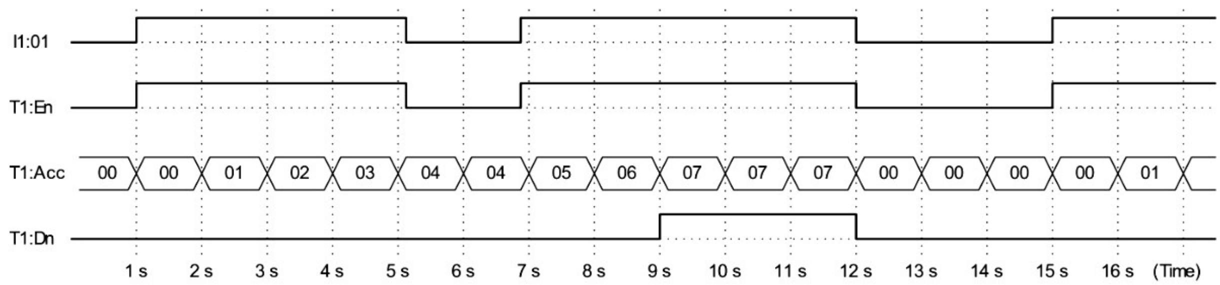
Retentive Timer (RTO) Instruction

The RTO instruction begins to count when its input rung conditions are true. The accumulated value is retained when the input rung conditions become false, and continues counting after the input rung conditions become true.

Timer ladder diagram example:



Typical timing diagram (Assume that Preset = 07)



Counter Instructions

Counter Instructions are output instructions used to count false-to-true rung transitions. These transitions are usually caused by events occurring at an input.

These counters can be UP (incrementing) or DOWN (decrementing).

Each counter instruction has two values (integers) associated with it:

- Accumulated Value (ACC): This is the current number of the counter. The initial value is zero.
- Preset Value (PR): This is a predetermined value set by the programmer. When the accumulated value is equal to, or greater than the preset value, a status bit is set. This bit can be used to control an output device.

Each counter is associated with two status bits:

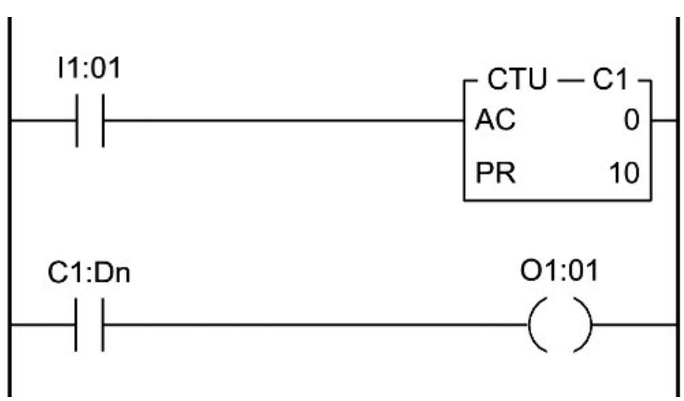
- Counter Enable Bit (EN): This bit is set when a false-to-true rung condition to the left of the counter instruction is detected.
- Done Bit (DN): This bit is set when the accumulated value is equal to the preset value. It is reset when the rung condition becomes false.

The maximum count value is 9999*. After a maximum count is reached, the counters reset and start counting from zero.

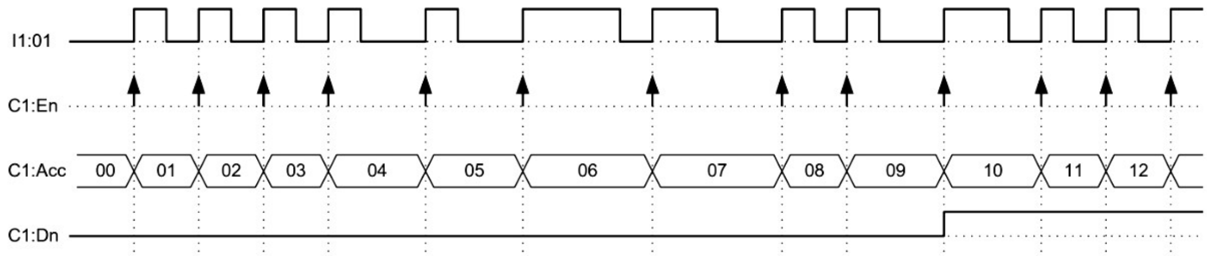
Count-up (CTU) Instruction

The CTU instruction increments its accumulated value on each false-to-true transition at its input, starting from 0.

Counter ladder diagram example:



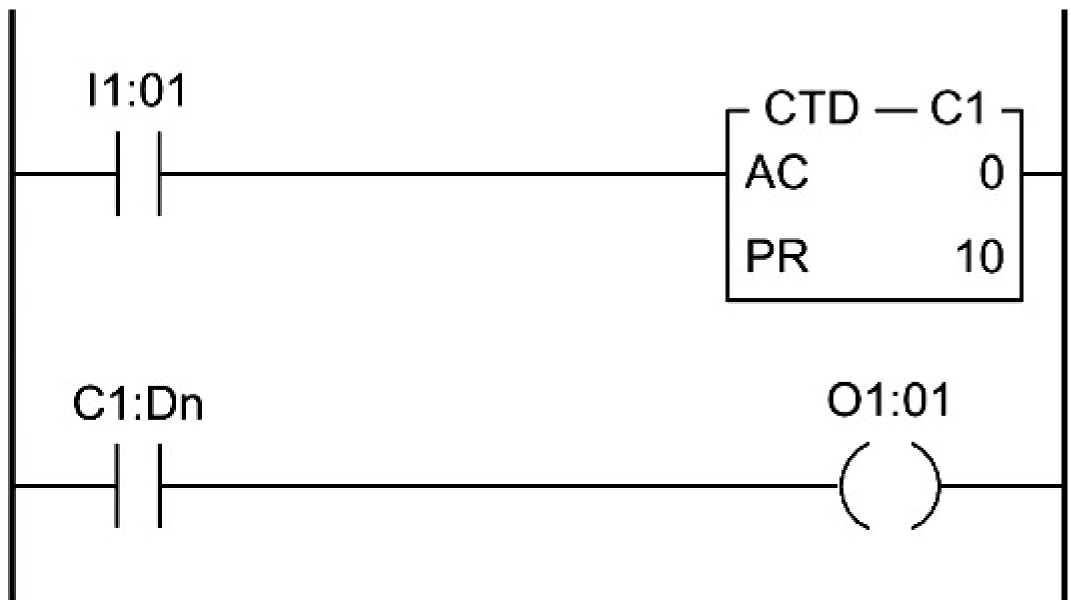
Typical timing diagram (Assume that Preset = 10)



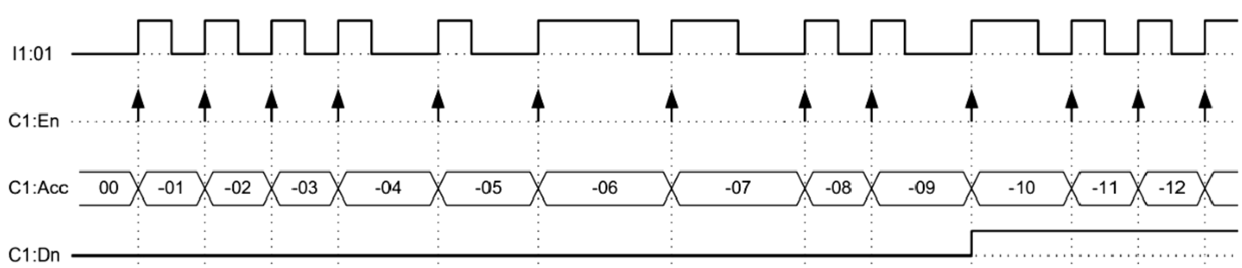
Count-down (CTD) Instruction

The CTD instruction decrements its accumulated value on each false-to-true transition at its input, starting from 0.

Counter ladder diagram example:



Typical timing diagram (Assume that Preset = -10)



The Reset (RES) Instruction

The RES instruction resets timing and counting instructions.

When the RES instruction is enabled it resets the following:

Counters:

- Accumulated value
- Counter Done Bit
- Counter Enabled Bit

Timers:

- Accumulated value
- Timer Done Bit
- Timer Timing Bit
- Timer Enable Bit

Reset ladder diagram example:

Some other PLC Instructions are :

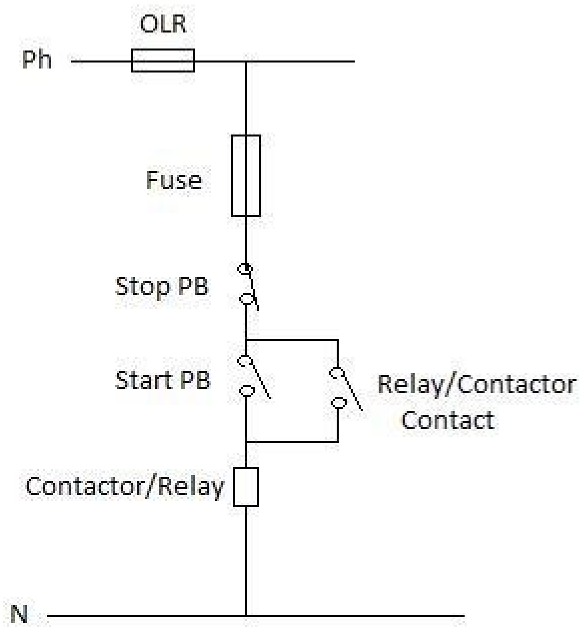
- Relay-type (Basic) instructions: I, O, OSR, SET, RES, T, C
- Data Handling Instructions:
- Data move Instructions: MOV, COP, FLL, TOD, FRD, DEG, RAD (degrees to radian).
- Comparison instructions: EQU (equal), NEQ (not equal), GEQ (greater than or equal), GRT (greater than).
- Mathematical instructions.
- Continuous Control Instructions (PID instructions).
- Program flow control instructions: MCR (master control reset), JMP, LBL, JSR, SBR, RET, SUS, REF
- Specific instructions:
- BSL, BSR (bit shift justify/right), SQO (sequencer output), SQC (sequencer compare), SQL (sequencer load).
- High-speed counter instructions: HSC, HSL, RES, HSE
- Communication instructions: MSQ, SVC
- ASCII instructions: ABL, ACB, ACI, ACL, CAN

LADDER DIAGRAMS FOR FOLLOWING (I) DOL STARTER AND STAR-DELTA STARTER (II) STAIR CASE LIGHTING (III) TRAFFIC LIGHT CONTROL (IV) TEMPERATURE CONTROLLER

(I) DOL STARTER AND STAR-DELTA STARTER

Working of Direct-On-Line (DOL) starter:

One method of starting electric motors is using direct on line (DOL) or across the line starter. In this method full line voltage is applied to the motor terminals. This is simplest type of motor starter. An electrical wiring diagram for single phase DOL starter is shown below.



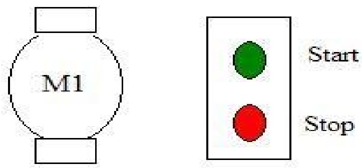
Control Circuit for DOL Starter

A DOL motor starter contains fuse and over load relay (OLR) for protection purpose. The starter can be contain momentary contact or maintained contact push buttons. The example considered here is momentary contact push buttons. For starting purpose normally open (NO) push button is preferred whereas normally closed (NC) push button is used to stop the motor.

The excessive supply voltage drop causing high inrush current is the criteria to limit the use of DOL starter. Conveyor motors, water pumps are the applications where DOL starters are used.

Procedure

Problem Statement: To start a motor using DOL starter. The simple P&I; diagram for this problem is as below.



Listing of Input and Output devices:

Inputs: PB1- To start the motor

PB2- To stop the motor

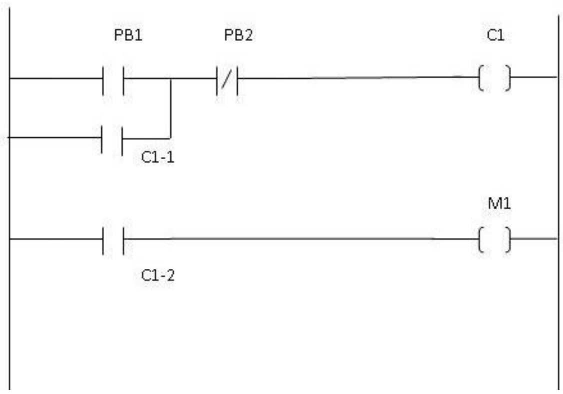
Output: M1- Motor

Sequence of Events :
1. When Start push button (PB1) is pressed, Motor (M1) has to start.
2. If Start pushbutton (PB1) is released and Stop pushbutton (PB2) is not pressed, Motor (M1) should remain on.
3. When Stop push button (PB2 is pressed, Motor (M1) has tol stop.
4. If stop push button is released and start is not pressed (released) motor should remain off.

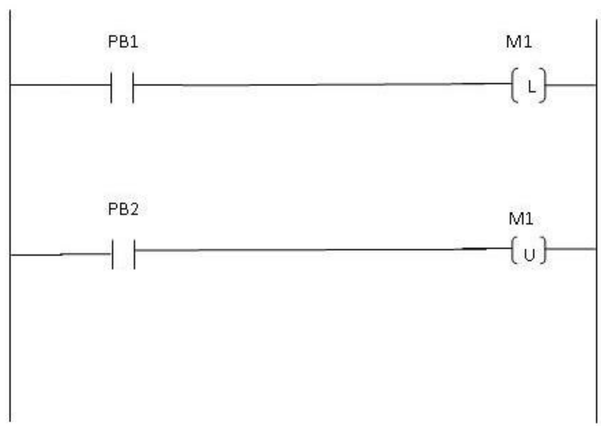
The Boolean equation to represent this sequence is

$$M1 = PB1.\overline{PB2}$$

The ladder diagram to implement these equations is shown below.



As the momentary contact push buttons are used here, the condition of PB1 is maintained through contact of coil C1. This contact is called as latching contact. The same sequence of event can be executed by using latch and unlatch instruction in the following way.



STAR-DELTA STARTER

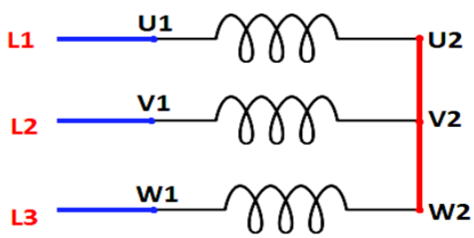
When electric motor is started, it draws a high current typical 5-6 times greater than normal current. In DC motors there is no back emf at starting therefore initial current is very high as compared to the normal current.

To protect the motor from these high starting currents we use a star and delta starter.

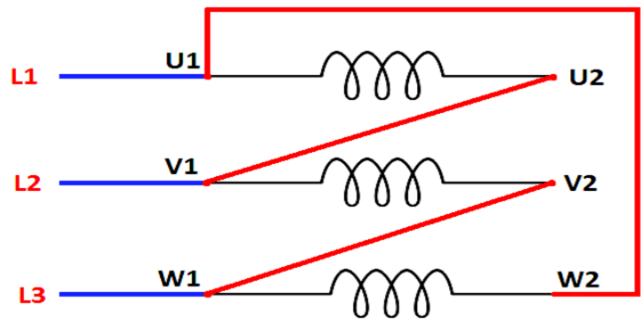
Simply in Star connection, supply voltage to motor will be less. so we use star connection during starting of the motor, after motor running we will change the connection form star to delta to gain full speed of the motor.

Star Delta Motor Starter

The following figure shows the winding connections in star and delta configuration one by one.

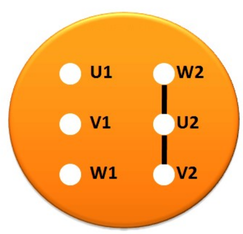


STAR CONNECTION

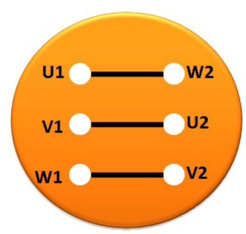


DELTA CONNECTION

It can be seen that in star connection, one end of all three windings are shorted to make star point while other end of each winding is connected to power supply.It can be seen that in star connection, one end of all three windings are shorted to make star point while other end of each winding is connected to power supply. In delta configuration, the windings are connected such that to make a close loop. The connection of each winding is shown in above figure. In actual motor the three phase connections are provided in the following order as shown

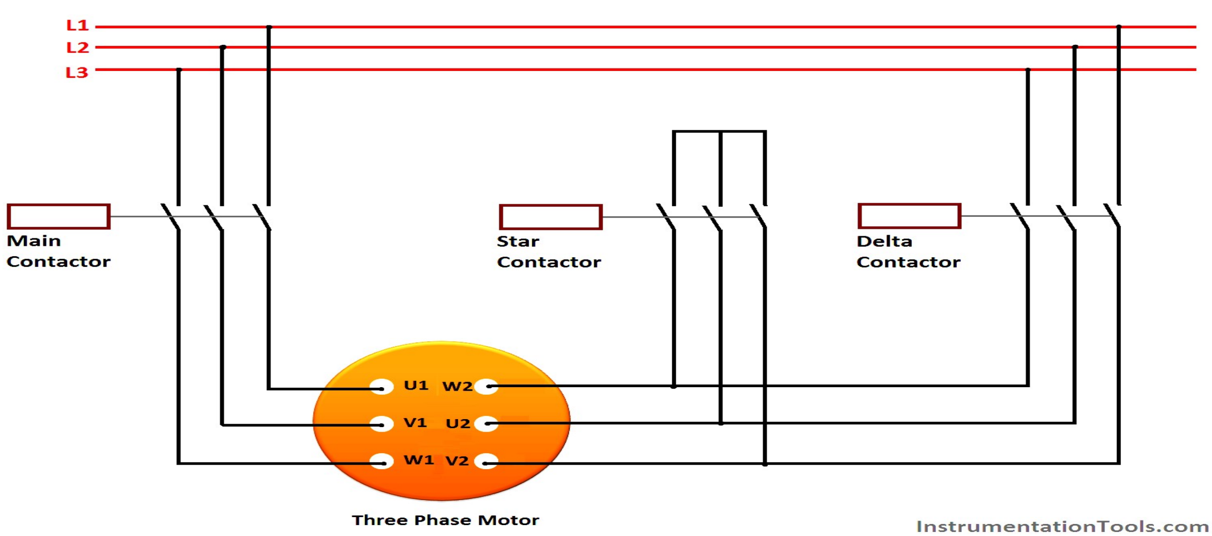


STAR CONNECTION

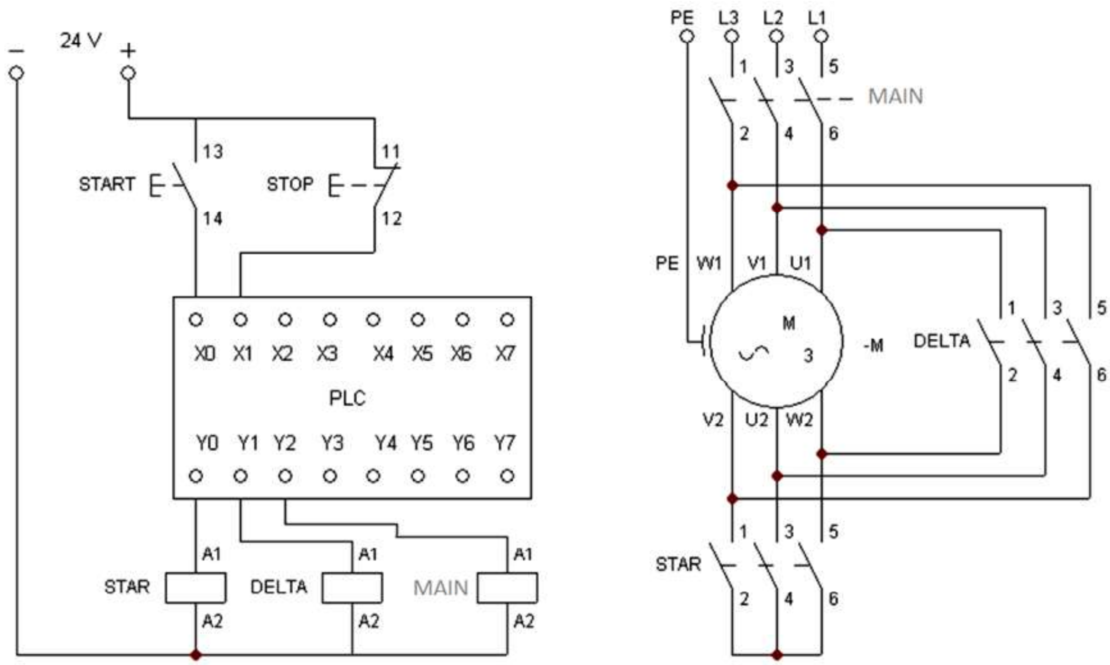


DELTA CONNECTION

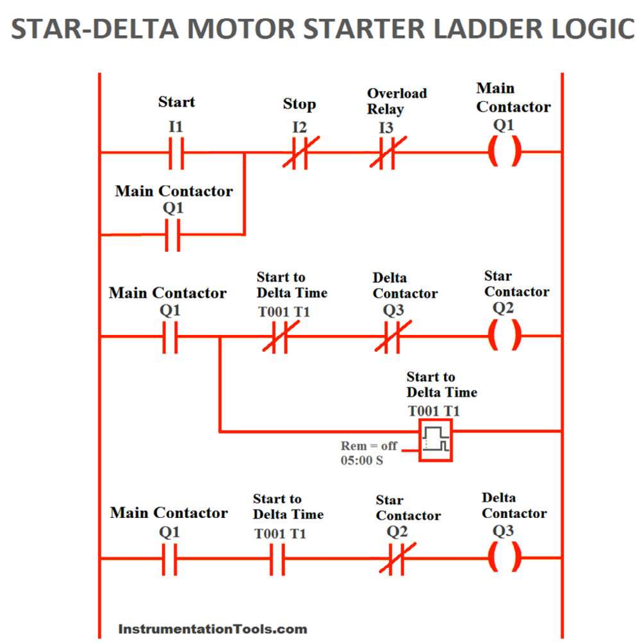
So in order to make winding connection in star and delta style in practical motor, the connection is shown above.



Main contractor is used to supply power to the windings. It must be turned on all the time. Initially the star contactor is closed while delta contactor is open It makes the motor windings in star configuration. When the motor gains speed, the star contactor is opened while delta contactor is closed turning the motor windings into delta configuration. The contactors are controlled by using PLC. The following section of PLC tutorial will explain the ladder programming for star delta motor starter.PLC program for star delta motor starter:



PLC Ladder Logic



Rung 1 Main contractor :

The main contactor depends upon the normally open input start push button (I1), normally closed stop button (I2) and normally closed overload relay. It means that Main contactor will only be energized if start button is pressed, while stop is not pressed and overload relay is not activated. A normally open input named (Q1) is added in parallel to the start button I1.By doing so, a push button is created which means that once motor is started, it will be kept started even if start button is released

Rung 2 Star contactor:

Star contactor depends upon main contractor, normally close contacts of timer (T1), and normally close contacts of output delta contactor (Q3).So star contactor will only be energized if main contactor is ON, time output is not activated and delta contactor is not energized.

Timer T1:

Timer T1 measures the time after which the winding connection of star delta starter is to be changed. It will start counting time after main contractor is energized.

Rung 3 Delta contactor:

Delta contactor will be energized when main contactor (Q1) is energized, timer T1 is activated and star contactor (Q3) is de-energized.

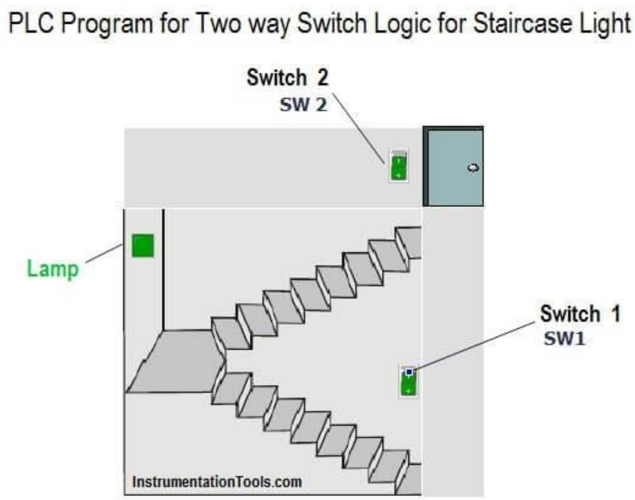
(ii) Stair case lighting

This is PLC Program for Two ways switch logic for staircase light in house

PLC Two Way Switch Logic

- In duplex type house there are ground floor and first floor and sometimes second floor also.
- Sometimes people need to go from ground floor to first floor or from first floor to ground floor by staircase provided in house.
- But in staircase there is no sunlight so people need a lamp/light to see the steps of the staircase easily.
- Here we are using a simple PLC to control this lamp using two switches, one switch at ground floor and second switch at first floor to control one lamp as shown in below figure.
- Note : we can also build the circuit using simple relays/switches also. This article only for understanding the basic concept of 2 way switch using a PLC Ladder Logic.

Problem Diagram



Problem Solution

We will solve this problem by simple automation. As shown in figure consider one simple house with one floor and staircase is provided in the house.

Here we will set lighting system for the users to switch ON/OFF the light whether they are on bottom of the stair or at top.

We will provide separate switch for each floor as shown in above figure.

List of inputs/outputs

Digital Inputs

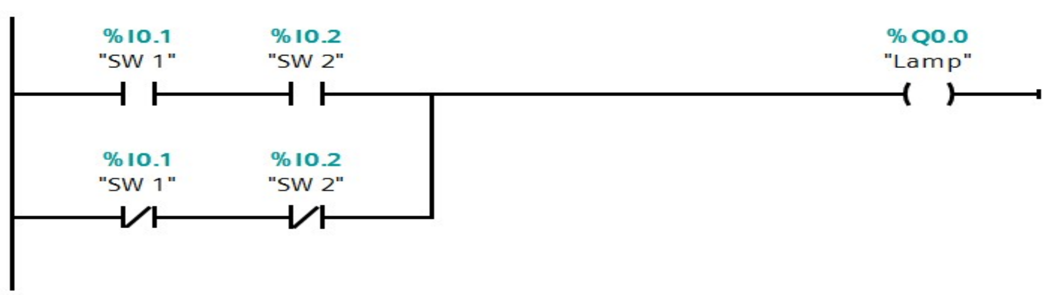
SW1: I0.1

SW2: I0.2

Digital Outputs

Lamp: Q0.0

PLC Ladder diagram for two ways switch logic



Program Description

- For this application, we used S7-1200 PLC and TIA portal software for programming.

- In above program, we have added two NO contacts of SW 1 (I0.1) and SW 2 (I0.2) in series and NC contacts of SW1 (I0.1) and SW2 (I0.2) in parallel of this series SW1 & SW2 NO Contacts.
- If the status of the bottom switch (SW1) and status of the top switch (SW2) are same then lamp will be ON. And if either status of the bottom or top switch is different from other then lamp (Q0.0) will be OFF.
- When lamp (Q0.0) is OFF then user can ON the lamp by changing status of any switch. Also user can turn OFF the lamp by changing the status of one of the two switches.

Runtime Test Cases

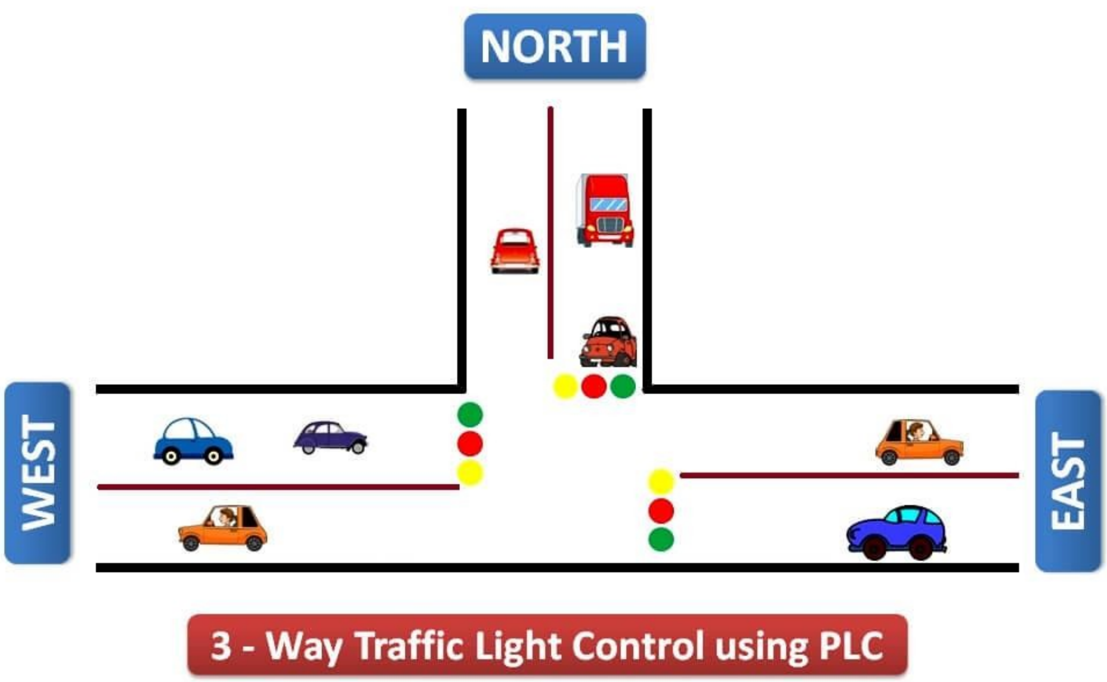
PLC Program for Two Way Switch Logic

Inputs	Outputs	Physical Elements
I0.1=1 & I0.2=1	Q0.0=1	Lamp on
I0.1=0 & I0.2=0	Q0.1=1	Lamp on
I0.1=0 & I0.2=1	Q0.1=0	Lamp off
I0.1=1 & I0.2=0	Q0.1=0	Lamp off

(iii) Traffic light Control

We most often come across a three-way traffic jam in our city. This **PLC program** gives the solution to control heavy traffic jams using programmable logic control.

Traffic Light Control using PLC



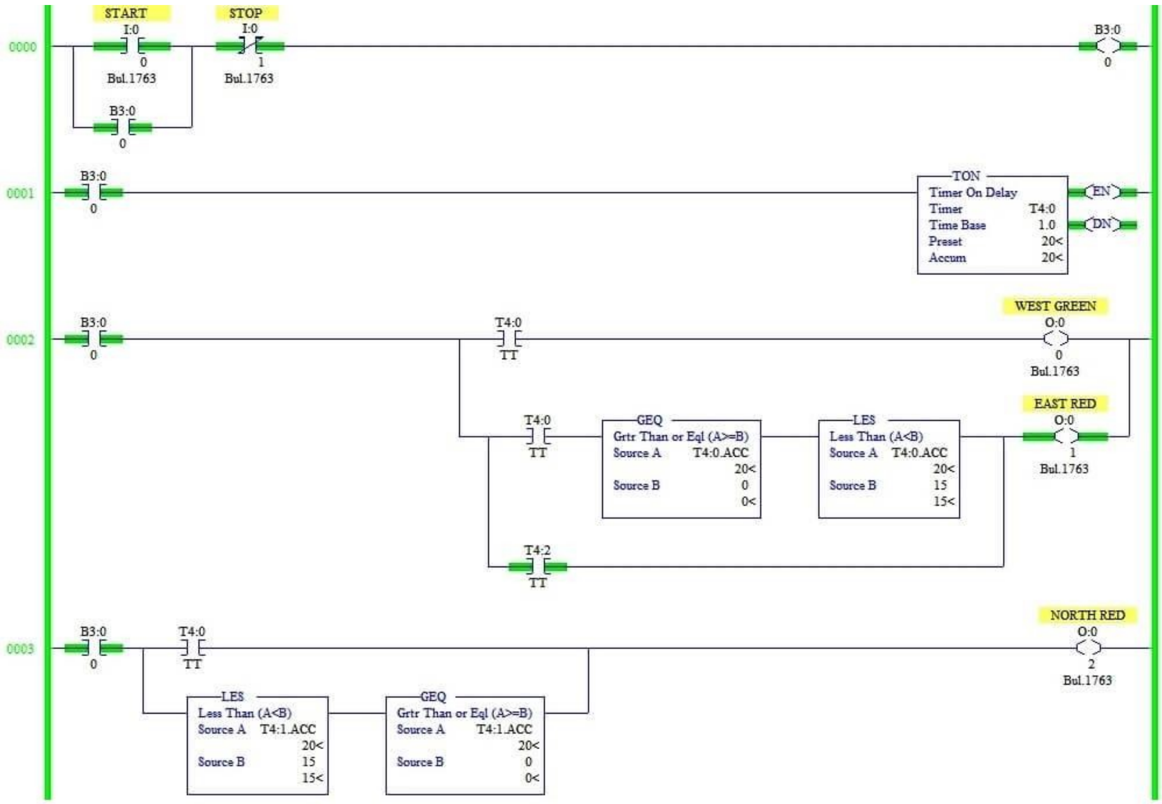
Problem Solution

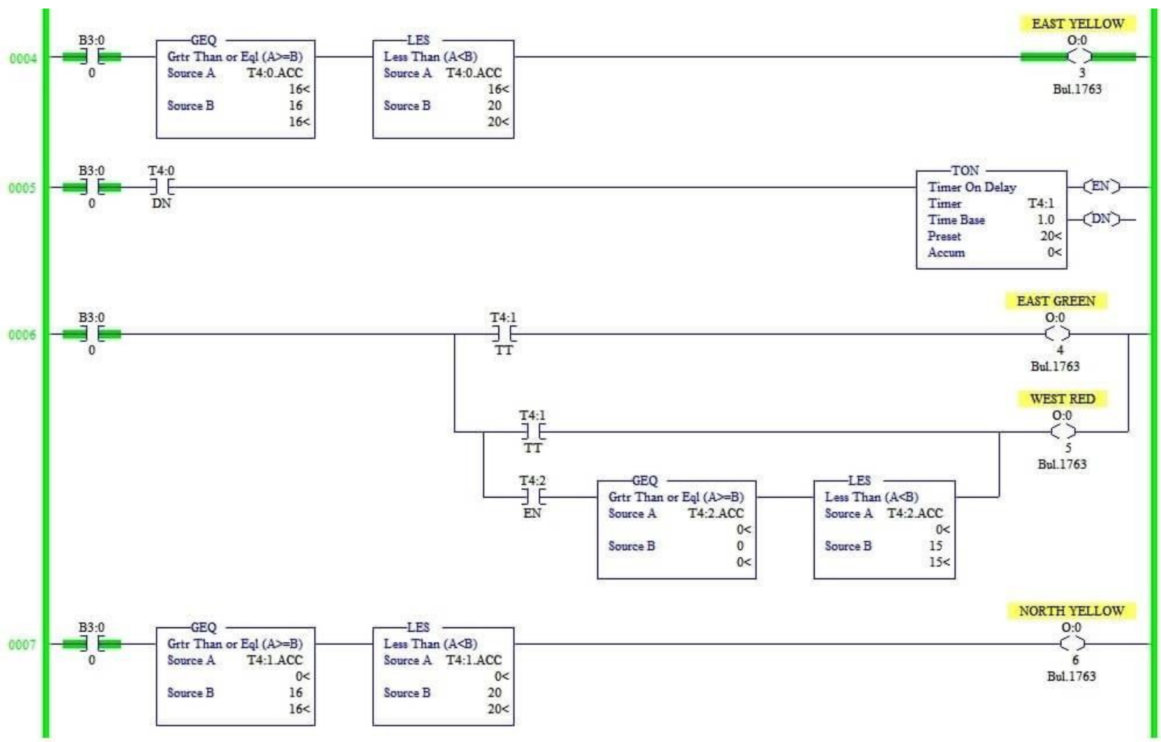
- They are so many ways to write a program for traffic light control ex: sequencer output method but in this normal input, outputs and timers are used.
- Timers are used to give time delay for output to turn ON and OFF.
- Reset coil is used at the end to run the program continuously.
- Comparator blocks are used to reduce the number of timers used.
- Program done in AB RSLogix 500 Software.

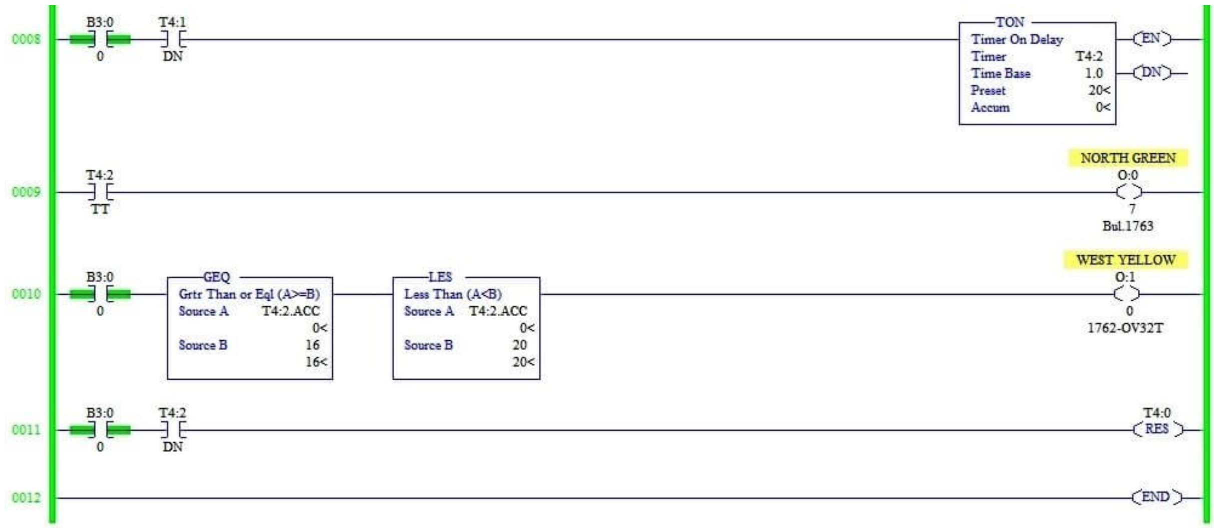
List of Inputs and Outputs for Traffic Control System

S.no	Address	Name	Input/Output
1	I:0/0	Start	Input
2	I:0/1	Stop	Input
3	B3.0	Memory	Memory
4	O:0/0	West Green	Output
5	O:0/1	East Red	Output
6	O:0/2	North Red	Output
7	O:0/3	East yellow	Output
8	O:0/4	East Green	Output
9	O:0/5	West Red	Output
10	O:0/6	North Yellow	Output
11	O:0/7	North Green	Output
12	O:1/0	West Yellow	Output

PLC Program for 3-way Traffic control System


0000
START
I:0
0
Bul.1763
STOP
I:0
1
Bul.1763
B3:0
0
0001
TON
Timer On Delay
Timer T4:0
Time Base 1.0
Preset 20<
Accum 20<
EN
DN
0002
T4:0
TT
WEST GREEN
O:0
0
Bul.1763
GEQ
Grtr Than or Eql (A>=B)
Source A T4:0.ACC
20<
Source B 0
0<
LES
Less Than (A<B)
Source A T4:0.ACC
20<
Source B 15
15<
EAST RED
O:0
1
Bul.1763
T4:2
TT
0003
NORTH RED
O:0
2
Bul.1763
LES
Less Than (A<B)
Source A T4:1.ACC
20<
Source B 15
15<
GEQ
Grtr Than or Eql (A>=B)
Source A T4:1.ACC
20<
Source B 0
0<


0004
B3:0
0
GEQ
Grtr Than or Eql (A>=B)
Source A T4:0.ACC
16<
Source B 16
16<
LES
Less Than (A<B)
Source A T4:0.ACC
16<
Source B 20
20<
EAST YELLOW
O:0
3
Bul.1763
0005
T4:0
DN
TON
Timer On Delay
Timer T4:1
Time Base 1.0
Preset 20<
Accum 0<
EN
DN
0006
T4:1
TT
EAST GREEN
O:0
4
Bul.1763
WEST RED
O:0
5
Bul.1763
T4:2
EN
GEQ
Grtr Than or Eql (A>=B)
Source A T4:2.ACC
0<
Source B 0
0<
LES
Less Than (A<B)
Source A T4:2.ACC
0<
Source B 15
15<
0007
GEQ
Grtr Than or Eql (A>=B)
Source A T4:1.ACC
0<
Source B 16
16<
LES
Less Than (A<B)
Source A T4:1.ACC
0<
Source B 20
20<
NORTH YELLOW
O:0
6
Bul.1763



Below tabular column gives the Steps or sequence of outputs to turn ON.

S.NO	EAST	WEST	NORTH
1	R	G	R
2	Y	G	R
3	G	R	R
4	G	R	Y
5	R	R	G
6	R	Y	G

PLC Logic Description for 3-way Traffic Control System

RUNG000:

This is a Latching rung to operate the system through Master Start and Stop PB.

RUNG001 and RUNG0002:

Starting the timer to turn ON first output West Green so east and west should be in red. Comparators in Parallel rung are used to turn OFF East red after 15 sec. Timer T4:2 timing bit in parallel contact used to turn ON East red again in 5$^{th\ and}$ 6Th Step. (Refer Above Tabular column for clarification)

RUNG 0003:

Turning ON North Red up to 3rd step using T4:0 and T4:1's timer timing bit and comparator blocks.

Rung 0004:

Turn ON East yellow for 5 sec using comparator blocks. (Step 2nd)

Rung 0005-0006-0007-0008-0009-0010 :

The same procedures followed to turn ON further outputs. (Refer Tabular column for a sequence of operation)

RUNG 0011:
Reset coil is turned ON using T4:2's done bit to restart the cycle from beginning
The program runs continuously until STOP PB is pressed

(IV) TEMPERATURE CONTROLLER

PLC Temperature Control : In a vessel there are Three Heaters which are used to control the temperature of the vessel.

PLC Temperature Control Programming

We are using Three Thermostats to measure the temperature at each heater. Also another thermostat for safety shutoff in case of malfunction or emergency or to avoid over temperatures. All these heaters have different set points or different temperature ranges where heaters can be turned ON accordingly (below table shows the temperature ranges).

- A temperature control system consists of four thermostats. The system operates three heating units. The thermostats (TS1/TS2/TS3/TS4 are set at 55°C, 60°C, 65°C and 70°C.
- Below 55°C temperature, three heaters (H1,H2,H3) are to be in ON state
- Between 55°C – 60°C two heaters (H2,H3) are to be in ON state.
- Between 60°C – 65°C one heater (H3) is to be in ON state.
- Above 70°C all heaters are to be in OFF state, there is a safety shutoff (Relay CR1) in case any heater is operating by mistake.
- A master switch turns the system ON and OFF.

Solution:

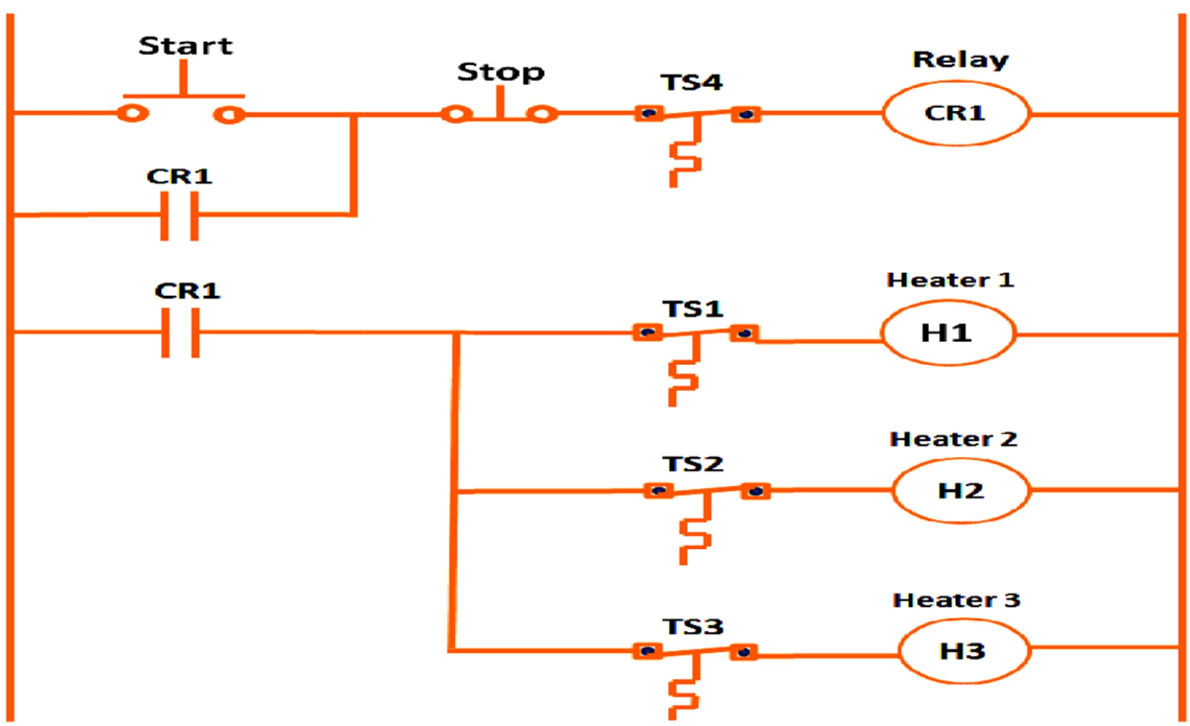
- There are four thermostats; assume them be in NC state when the set point is not reached.
- Let there be a control relay (CR1) to work as a safety shutoff.
- Master Switch: The Start switch is NO and Stop switch NC type.

The below table shows the temperature ranges where Thermostats (TS1,TS2,TS3,TS4) status will be indicated as per the temperature value.

Also the Heaters (H1, H2, H3) status in which either those Heaters will be ON or OFF as per the temperature value.

Temperature	Thermostats	Heater 1	Heater 2	Heater3
Below 55°C	TS1 Closed TS2 Closed TS3 Closed TS4 Closed	ON	ON	ON
55°C-60°C	TS1 Open TS2 Closed TS3 Closed TS4 Closed	OFF	ON	ON
60°C-65°C	TS1 Open TS2 Open TS3 Closed TS4 Closed	OFF	OFF	ON
65°C-70°C	TS1 Open TS2 Open TS3 Open TS4 Closed	OFF	OFF	OFF
Above 70°C	TS1 Open TS2 Open TS3 Open TS4 Open	OFF	OFF	OFF

PLC Ladder Logic

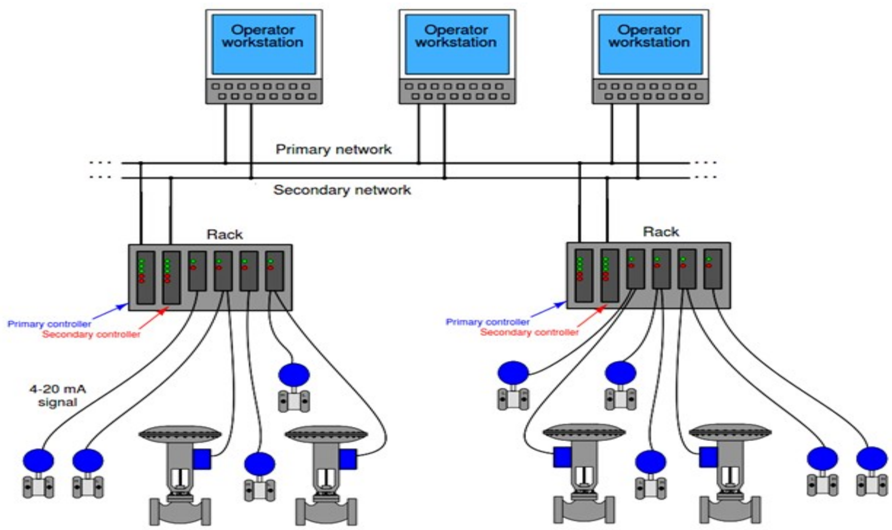


Ladder Logic Operation :
First Rung:
It has START button (default NO contact) and STOP button (default NC contact). A Relay CR1 is used to control the heaters depending on the thermostats status.
A Thermostat TS4 is connected in between STOP & Relay; if TS4 activated (means TS4 contact changes from NC to NO) then all heaters will be OFF.
An NO contact of Relay CR1 is used across the START button in order to latch or hold the START command.
Second Rung:
Second Rung:
An NO contact of Relay CR1 is used to control the Heaters (H1, H2, H3) with the thermostats (TS1,TS2,TS3) status.
After giving START command, This NO contact becomes NC contact. if temperature below 55 Deg C, TS1, TS2 & TS3 will be in Close Status so all heaters will be ON.
If Temperature is in between 55 to 60 Deg C, Then TS1 will be Open, so Heater H1 will be OFF.
Then, if temperature in between 60 to 65 Deg C then TS2 also be Open, so Heater H2 will be OFF
if temperature in between 65 to 70 Deg C then TS3 also be Open, so Heater H3 will be OFF
There is a safety Shutoff which is used to avoid any malfunctions of Thermostats or to avoid over temperatures.
if temperature reaches above 70 Deg C then TS4 will activates and de-energizes the Relay, thus all Heaters will be turned OFF.
Note: Here Heaters H1, H2, H3 are either Relays or Contactors we energizing. so an NO contact of these relays are connected to Electrical Heater feeder circuits (MCC).These Electrical Feeder circuits will be controlled as per these signals and accordingly the heaters will be either ON or OFF.

SPECIAL CONTROL SYSTEMS- BASICS DCS & SCADA SYSTEMS

Distributed Control Systems (DCS)

The following illustration shows a typical distributed control system (DCS) architecture:



Each "rack" contains a microprocessor to implement all necessary control functions, with individual I/O (input/output) "cards" for converting analog field instrument signals into digital format, and vice-versa. Redundant processors, redundant network cables, and even redundant I/O cards address the possibility of component failure.

DCS processors are usually programmed to perform routine self-checks on redundant system components to ensure availability of the spare components in the event of a failure.

If there ever was a total failure in one of the "control racks" where the redundancy proved insufficient for the fault(s), the only PID loops faulted will be those resident in that rack, not any of the other loops throughout the system.

Likewise, if ever the network cables become severed or otherwise faulted, only the information flow between those two points will suffer; the rest of the system will continue to communicate data normally.

Thus, one of the "hallmark" features of a DCS is its tolerance to serious faults: even in the event of severe hardware or software faults, the impact to process control is minimized by design.

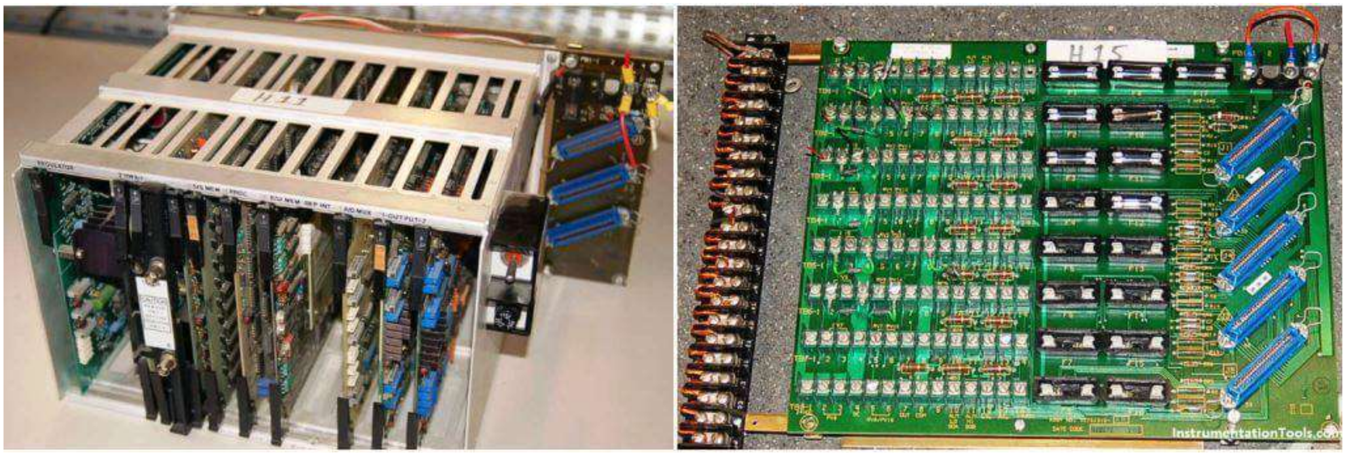
One of the very first distributed control systems in the world was the Honeywell TDC2000 system (Note 1) , introduced in 1975. By today's standards, the technology was crude, but the concept was revolutionary.

Note 1: To be fair, the Yokogawa Electric Corporation of Japan introduced their CENTUM distributed control system the same year as Honeywell.

Each rack (called a "box" by Honeywell) consisted of an aluminum frame holding several large printed circuit boards with card-edge connectors. A "basic controller" box appears in the left-hand photograph.

The right-hand photograph shows the termination board where the field wiring (420 mA) connections were made. A thick cable connected each termination board to its respective controller box:

DCS Hardware



Controller redundancy in the TDC2000 DCS took the form of a "spare" controller box serving as a backup for up to eight other controller boxes. Thick cables routed all analog signals to this spare controller, so that it would have access to them in the event it needed to take over for a failed controller.

The spare controller would become active on the event of any fault in any of the other controllers, including failures in the I/O cards.

Thus, this redundancy system provided for processor failures as well as I/O failures. All TDC2000 controllers communicated digitally by means of a dual coaxial cable network known as the "Data Hiway." The dual cables provided redundancy in network communications.

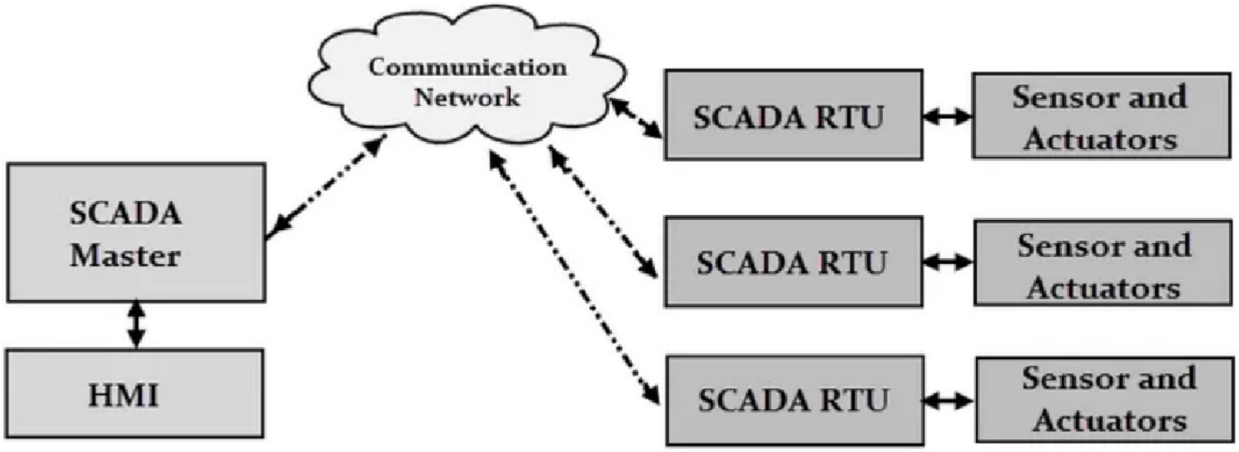
SCADA SYSTEMS

SCADA stands for "Supervisory Control and Data Acquisition". SCADA is a type of process control system architecture that uses computers, networked data communications and graphical Human Machine Interfaces (HMIs) to enable a high-level process supervisory management and control.

SCADA systems communicate with other devices such as programmable logic controllers (PLCs) and PID controllers to interact with industrial process plant and equipment.

SCADA systems form a large part of control systems engineering. SCADA systems gather pieces of information and data from a process that is analyzed in real-time (the "DA" in SCADA). It records and logs the data, as well as representing the collected data on various HMIs.

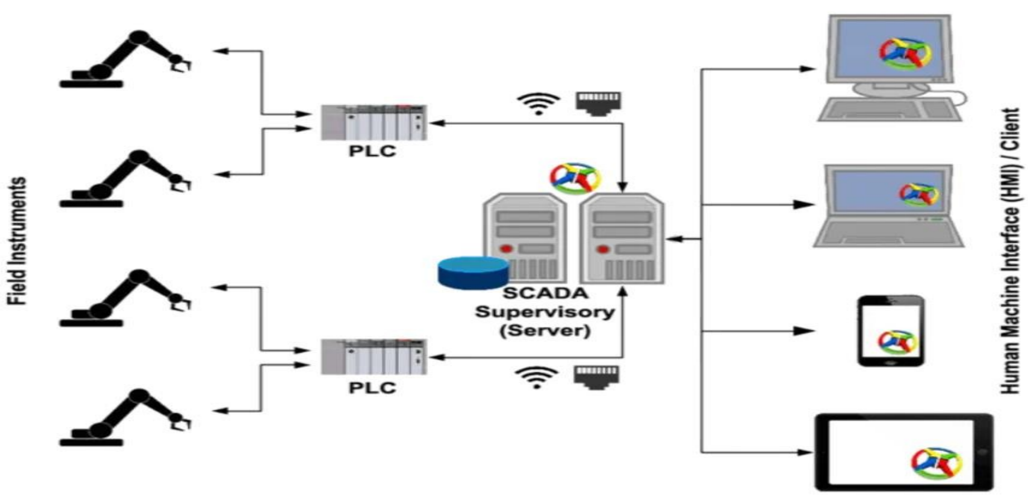
This enables process control operators to supervise (the "S" in SCADA) what is going on in the field, even from a distant location. It also enables operators to control (the "C" in SCADA) these processes by interacting with the HMI.



Supervisory Control and Data Acquisition systems are essential to a wide range of industries and are broadly used for the controlling and monitoring of a process. SCADA systems are prominently used as they have the power to control, monitor, and transmit data in a smart and seamless way. In today's data-driven world, we are always looking for ways to increase automation and make smarter decisions through the proper use of data – and SCADA systems are a great way of achieving this.

SCADA systems can be run virtually, which allows the operator to keep a track of the entire process from his place or control room. Time can be saved by using SCADA efficiently. One such excellent example is, SCADA systems are used extensively in the Oil and Gas sector. Large pipelines will be used to transfer oil and chemicals inside the manufacturing unit.

Hence, safety plays a crucial role, such that there should not be any leakage along the pipeline. In case, if some leakage occurs, a SCADA system is used to identify the leakage. It infers the information, transmits it to the system, displays the information on the computer screen and also gives an alert to the operator.



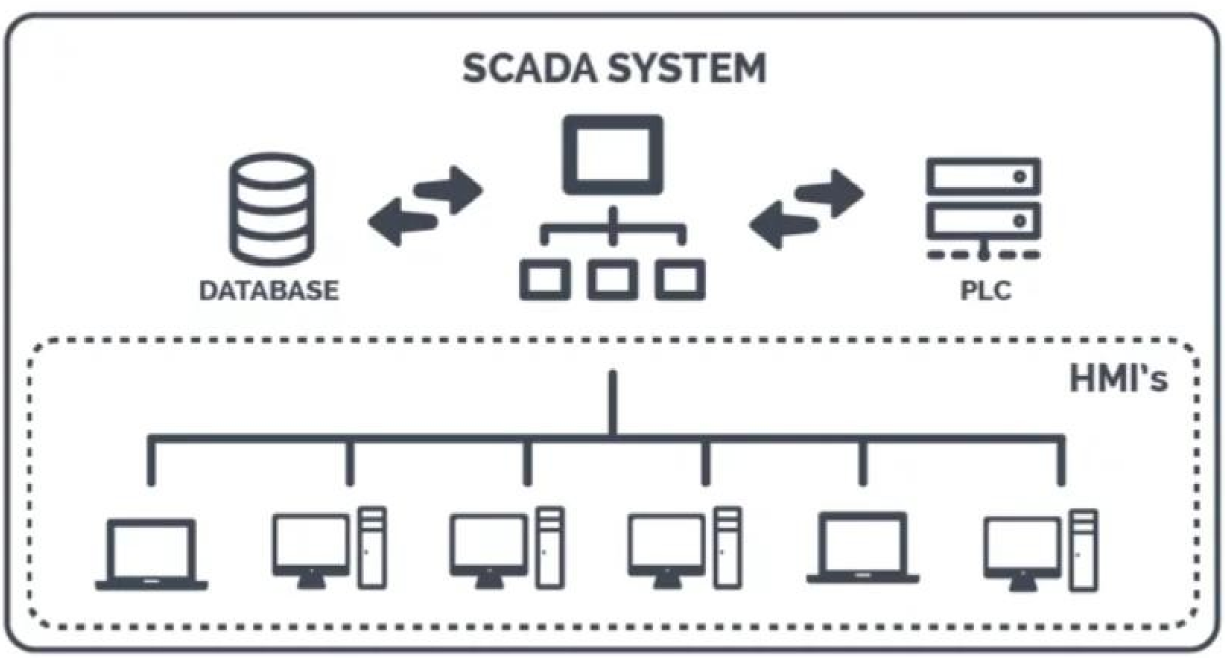
SCADA Architecture

Generic SCADA systems contain both hardware and software components. The computer used for analysis should be loaded with SCADA software. The hardware component receives the input data and feds it into the system for further analysis.
SCADA system contains a hard disk, which records and stores the data into a file, after which it is printed as when needed by the human operator. SCADA systems are used in various industries and manufacturing units like Energy, Food and Beverage, Oil and Gas, Power, Water, and Waste Management units and many more.
SCADA Basics
Objectives of SCADA

- Monitor: SCADA systems continuously monitor the physical parameters
- Measure: It measures the parameter for processing
- Data Acquisition: It acquires data from RTUs (Remote Terminal Units), data loggers, etc
- Data Communication: It helps to communicate and transmit a large amount of data between MTU and RTU units
- Controlling: Online real-time monitoring and controlling of the process
- Automation: It helps for automatic transmission and functionality

The SCADA systems consist of hardware units and software units. SCADA applications are run using a server. Desktop computers and screens act as an HMI which are connected to the server. The major components of a SCADA system include:

- Master Terminal Unit (MTU)
- Remote Terminal Unit (RTU)
- Communication Network (defined by its network topology)



Master Terminal Unit (MTU)
MTU is the core of the SCADA system. It comprises a computer, PLC and a network server that helps MTU to communicate with the RTUs. MTU begins communication, collects and saves data, helps to interface with operators and to communicate data to other systems.

Remote Terminal Unit (RTU)
Being employed in the field sites, each Remote Terminal Unit (RTU) is connected with sensors and actuators. RTU is used to collect information from these sensors and further sends the data to MTU. RTUs have the storage capacity facility.

So, it stores the data and transmits the data when MTU sends the corresponding command. Recently developed units are employed with sophisticated systems that utilize PLCs as RTUs. This helps for direct transfer and control of data without any signal from MTU.

Remote terminal unit

Communication Network
In general, network means connection. When you tell a communication network, it is defined as a link between RTU in the field to MTU in the central location. The bidirectional wired or wireless communication channel is used for networking purposes. Various other communication mediums like fiber optic cables, twisted pair cables, etc. are also used.
Functions of SCADA Systems
In a nutshell, we can tell the SCADA system is a collection of hardware and software components that allows the manufacturing units to perform specific functions. Some of the important functions include

- To monitor and gather data in real-time
- To interact with field devices and control stations via Human Machine Interface (HMI)
- To record systems events into a log file
- To control manufacturing process virtually
- Information Storage and Reports

COMPUTER CONTROL–DATA ACQUISITION, DIRECT DIGITAL CONTROL SYSTEM (BASICS ONLY)

DATA ACQUISITION SYSTEMS

Definition

Data acquisition is the process of real world physical conditions and conversion of the resulting samples into digital numeric values that can be manipulated by a computer. Data acquisition and data acquisition systems (abbreviated with the acronym DAS) typically involves the conversion of analog waveforms into digital values for processing.

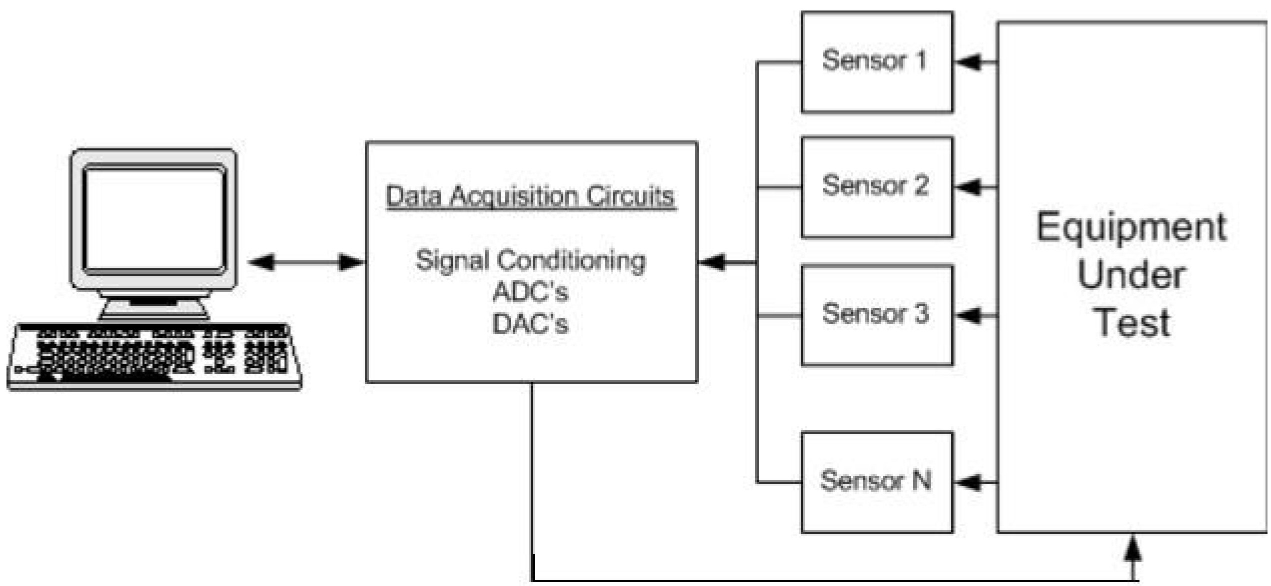
The components of data acquisition systems include:

i) Sensors that convert physical parameters to electrical signals.

ii) Signal conditioning circuitry to convert sensor signals into a form that can be converted to digital values.

iii) Analog-to-digital converters, which convert conditioned sensor signals to digital values.

Diagram



Fundamental elements of data acquisition system

Explanation

Data acquisition is the process of extracting, transforming, and transporting data from the source systems and external data sources to the data processing system to be displayed, analyzed, and stored.

A data acquisition system (DAQ) typically consist of transducers for asserting and measuring electrical signals, signal conditioning logic to perform amplification, isolation, and filtering, and other hardware for receiving analog signals and providing them to a processing system, such as a personal computer.

Data acquisition systems are used to perform a variety of functions, including laboratory research, process monitoring and control, data logging, analytical chemistry, tests and analysis of physical phenomena, and control of mechanical or electrical machinery.
Data recorders are used in a wide variety of applications for imprinting various types of forms, and documents.

Data collection systems or data loggers generally include memory chips or strip charts for electronic recording, probes or sensors which measure product environmental parameters and are connected to the data logger.
Hand-held portable data collection systems permit in field data collection for up-to-date information processing.

Source
Data acquisition begins with the physical phenomenon or physical property to be measured.
Examples of this include temperature, light intensity, gas pressure, fluid flow, and force.
Regardless of the type of physical property to be measured, the physical state that is to be measured must first be transformed into a unified form that can be sampled by a data acquisition system.
The task of performing such transformations falls on devices called sensors.
A sensor, which is a type of transducer, is a device that converts a physical property into a corresponding electrical signal (e.g., a voltage or current) or, in many cases, into a corresponding electrical characteristic (e.g., resistance or capacitance) that can easily be converted to electrical signal.

The ability of a data acquisition system to measure differing properties depends on having sensors that are suited to detect the various properties to be measured. There are specific sensors for many different applications.
DAQ systems also employ various signal conditioning techniques to adequately modify various different electrical signals into voltage that can then be digitized using an Analog-to-digital converter (ADC).

Signals
Signals may be digital (also called logic signals sometimes) or analog depending on the transducer used. Signal conditioning may be necessary if the signal from the transducer is not suitable for the DAQ hardware being used.
The signal may need to be amplified, filtered or demodulated.
Various other examples of signal conditioning might be bridge completion, providing current or voltage excitation to the sensor, isolation, and linearization. For transmission purposes, single

ended analog signals, which are more susceptible to noise can be converted to differential signals. Once digitized, the signal can be encoded to reduce and correct transmission errors.

DAQ hardware

DAQ hardware is what usually interfaces between the signal and a PC. It could be in the form of modules that can be connected to the computer's ports (parallel, serial, USB, etc.) or cards connected to slots (S-100 bus, Apple Bus, ISA, MCA, PCI, PCI-E, etc.) in the mother board.

Usually the space on the back of a PCI card is too small for all the connections needed, so an external breakout box is required. The cable between this box and the PC can be expensive due to the many wires, and the required shielding

DAQ cards often contain multiple components (multiplexer, ADC, DAC, TTL-IO, high speed timers, RAM). These are accessible via a bus by a microcontroller, which can run small programs.

A controller is more flexible than a hard wired logic, yet cheaper than a CPU so that it is alright to block it with simple polling loops.

The fixed connection with the PC allows for comfortable compilation and debugging. Using an external housing a modular design with slots in a bus can grow with the needs of the user.

Not all DAQ hardware has to run permanently connected to a PC, for example intelligent stand-alone loggers and oscilloscopes, which can be operated from a PC, yet they can operate completely independent of the PC.

DAQ software

DAQ software is needed in order for the DAQ hardware to work with a PC. The device driver performs low-level register writes and reads on the hardware, while exposing a standard API for developing user applications.

A standard API such as COMEDI allows the same user applications to run on different operating systems, e.g. a user application that runs on Windows will also run on Linux and BSD.

Advantages

Reduced data redundancy

Reduced updating errors and increased consistency

Greater data integrity and independence from applications programs

Improved data access to users through use of host and query languages

Improved data security

Reduced data entry, storage, and retrieval costs

Facilitated development of new applications program

Disadvantages

Database systems are complex, difficult, and time-consuming to design Substantial hardware and software start-up costs

Damage to database affects virtually all applications programs

Extensive conversion costs in moving form a file-based system to a database system Initial training required for all programmers and users

Applications

Temperature measurement

Recommended application software packages and necessary toolkit

Prewritten Lab VIEW example code, available for download

Sensor recommendations

DIRECT DIGITAL CONTROL SYSTEM

Direct Digital Control (DDC) is the automated control of a condition or process by a digital device.

DDC takes a centralized network-oriented approach. All instrumentation is gathered by various analog and digital converters which use the network to transport these signals to the central controller.

The centralized computer then follows all of its production rules (which may incorporate sense points anywhere in the structure) and causes actions to be sent via the same network to valves, actuators, and other HVAC components that can be adjusted.

A microprocessor operating at sufficient clock speed is able to execute more than one PID control algorithm for a process loop, by "time-sharing" its calculating power: devoting slices of time to the evaluation of each PID equation in rapid succession.

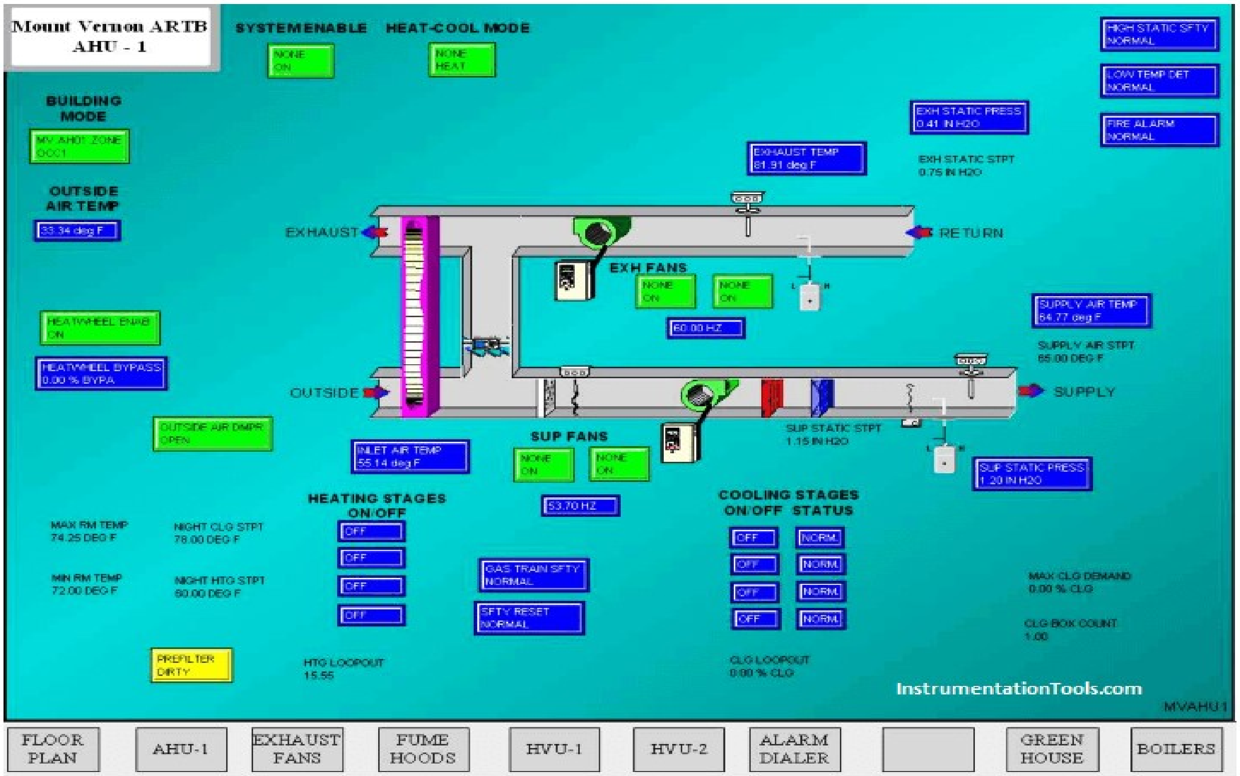
This not only makes multiple-loop digital control possible for a single microprocessor, but also makes it very attractive given the microprocessor's natural ability to manage data archival, transfer, and networking.

A single computer is able to execute PID control for multiple loops, and also make that loop control data accessible between loops (for purposes of cascade, ratio, feed forward, and other control strategies) and accessible on networks for human operators and technicians to easily access.

Such direct digital control (DDC) has been applied with great success to the problem of building automation, where temperature and humidity controls for large structures benefit from large-scale data integration.

Operators, engineers, and technicians alike must use software running on a networked personal computer to access data in this control system.

An example of the HMI (Human-Machine Interface) software one might see used in conjunction with a DDC controller is shown here, also from a Siemens APOGEE building control system:



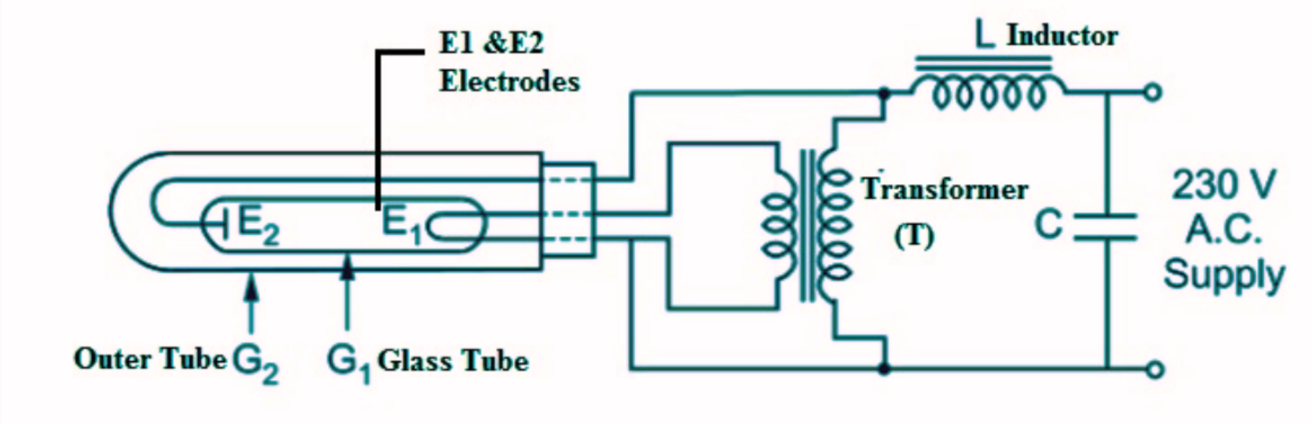
This particular screenshot shows monitored and controlled variables for a heat exchanger ("heat wheel") used to exchange heat between outgoing and incoming air for the building.

5TH ELECTRICAL

GANDHI ACADEMY OF TECHNOLOGY AND ENGINEERING

(Approved By AICTE New Delhi & Affiliated to BPUT, Rourkela, Odisha)

UTILIZATION OF ELECTRICAL ENERGY & TRACTION



Prepared By: Susanta Kumar Sahu

As Per SCTE&VT , Odisha Syllabus

CHAPTER-1
ELECTROLYSIS

Definition and Basic principle of Electro Deposition.

Electro deposition is the process of coating a thin layer of one metal on top of different metal to modify its surface properties. It is done to achieve the desire electrical and corrosion resistance, reduce wear &friction, improve heat tolerance and for decoration.

Electroplating Basics

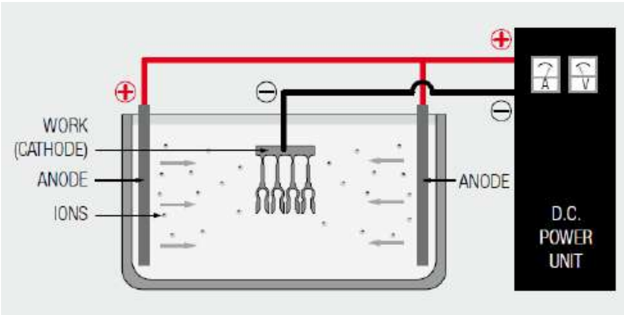


Fig-1. Electrochemical Plating

Figure- 1, schematically illustrates a simple electrochemical plating system. The "electro" part of the system includes the voltage/current source and the electrodes, anode and cathode, immersed in the "chemical" part of the system, the electrolyte or plating bath, with the circuit being completed by the flow of ions from the plating bath to the electrodes. The metal to be deposited may be the anode and be ionized and go into solution in the electrolyte, or come from the composition of the plating bath. Copper, tin, silver and nickel metal usually comes from anodes, while gold salts are usually added to the plating bath in a controlled process to maintain the composition of the bath. The plating bath generally contains other ions to facilitate current flow between the electrodes. The deposition of metal takes place at the cathode. The overall plating process occurs in the following sequence:

1. Power supply pumps electrons into the cathode.
2. An electron from the cathode transfers to a positively charged metal ion in the solution and the reduced metal plates onto the cathode.
3. Ionic conduction through the plating bath completes the circuit to the anode.
4. At the anode two different processes take place depending on whether the anode material is soluble, the source of the metal to be plated, or insoluble, inert. If the anode material is soluble, a metal atom gives up an electron and goes into the solution as a positively charged metal ion replenishing the metal content of the plating bath. If the anode is inert a negatively charged ion from the plating bath gives up an electron to the anode.
5. The electron flows from the anode to the power supply completing the circuit. The deposition of metal at the cathode requires an electron so the rate of deposition

depends on the flow of electrons, that is, the current flowing from the rectifier. The thickness of the deposit, therefore, depends on the current and the length of time the current is applied. This relationship is a result of Faraday's law which relates the weight of a substance produced by an anode or cathode electrode reaction during electrolysis as being directly proportional to the quantity of electricity passed through the cell.

Faraday's Laws of Electrolysis

From his experiments, Faraday deduced two fundamental laws which govern the phenomenon of electrolysis. These are:

(i) **First Law.** The mass of ions liberated at an electrode is directly proportional to the quantity of electricity i.e. charge which passes through the electrolyte.

Or

The weight of a substance liberated from an electrolyte in a given time is proportional to the quantity of electricity passing through the electrolyte.

That is $W \propto Q \propto It$, where I is the current and t is the time.

$W = Zit$

Where Z is a constant called electro-chemical equivalent.

If I = 1 ampere and T= one second then,

$Z = W$, which gives a definition of Z.

The electro-chemical equivalent of a substance is the amount of that substance by weight liberated in unit time by unit current.

(ii) **Second Law.** The masses of ions of different substances liberated by the same quantity of electricity are proportional to their chemical equivalent weights.

or,

If the same current flows through several electrolytes, the weights of ions liberated are proportional to their chemical equivalents.

The chemical equivalent of a substance is the weight of the substance which can displace or combine with unit weight of hydrogen .Obviously, the chemical equivalent of hydrogen is 1 by definition.

DEFINITIONS

1. **Current Efficiency**

On account of the impurities which cause secondary reactions, the quantity of a substance liberated is less than that calculated from faraday's Law.

Current efficiency is the ratio of the actual mass of a substance liberated from an electrolyte by the passage of current to the theoretical mass liberated according to Faraday's law. Current efficiency can be used in measuring electro deposition thickness on materials in electrolysis. Current efficiency is also known as faradic efficiency, faradic yield and columbic efficiency.

2. Energy Efficiency

On account of secondary reactions, the voltage actually required for the deposition or liberation of metal is higher than the theoretical value which increases the actual energy required.

Energy efficiency is defined as $=\frac{\text{theoretical energy}}{\text{actual energy required}}$

It is a process by which a metal is deposited over another metal or non-metal. Electro-plating is a very common example of such process.

Conditions have to be provided so that the deposit will be fine grained and will have a smooth appearance. The factors which affect the electro-deposition of metals are :

(i) **Current Density**

(ii) **Electrolyte concentration**

(iii) **Temperature**

(iv) **Addition agents**

(v) **Nature of electrolyte**

(vi) **Nature of the metal on which the deposit is to be made**

(vii) **Throwing power of the electrolyte**

Current density

At low values of current density the ions are released at a slow rate and the rate of growth of nuclei is more than the rate at which the new nuclei form themselves. Electro-deposition depends upon the rate at which crystals grow and the rate at which fresh nuclei are formed. Therefore, at low current densities the deposit will be coarse and crystalline in nature. At higher values of current density the quality of deposit becomes more uniform and fine-grained on account of the greater rate of formation of nuclei. If the current density is so high that it exceeds the limiting value for the electrolyte hydrogen is released and spongy and porous deposit is obtained.

Electrolytic Concentration

This is more or less complementary to the first factor, i.e. current density, since by increasing the concentration of the electrolyte higher current density can be achieved. Increase of concentration tends to give better deposits and some people therefore favour it.

Temperature

The temperature of the electrolyte has two contradictory effects. One, at comparatively high temperature there is more diffusion and even at relatively high current density smooth deposits may be produced. Two, the rate of crystal growth increases the possibility of coarse deposits. At moderate temperatures the deposits are good. In chromium plating the temperature is maintained at 35^0C, and in nickel between 50^0C to 60^0C .

Addition Agents

the quality of a deposit is improved by the presence of an addition agent which may be colloidal matter or an organic compound, otherwise the metal deposits in the form of large crystals and the surface becomes rough. Materials used as addition agents are gelatin, agar, glue, gums, rubber, alkaloids, sugar etc. The addition agents are supposed to be absorbed by crystal nuclei and prevent their growth into large crystals. The discharged ions start to build up new nuclei and the deposit of metal is fine-grained.

Nature of electrolyte

Smooth deposits are obtained from solutions having complex ions, e.g.,cyanides. Silver from nitrate solution forms a coarse deposit while from cyanide solution it forms a smooth deposit. Therefore, the formation of smooth deposit largely depends upon the nature of electrolyte used.

Nature of the metal on which deposit is to be made

This factor influences the growth of crystals since it is believed that the operation of crystals is in continuation of these in the base metal.

Throwing Power

The throwing power of an electrolyte may be regarded as the quality which produces a uniform deposit on a cathode having an irregular shape.Since the shape is irregular, The distance of the various parts of the cathode from the anode is not the same and therefore the conductance of the electrolyte is not the same for all parts of the cathode. The phenomenon of throwing power has not been clearly understood so far. In an electrolyte of low conductance, the current will concentrate on the parts of the cathode which are nearer the cathode resulting in poor throwing power. If the electrolyte has good conductance, the throwing power will also be good. One way to improve the throwing power is to keep a good distance between the cathode and the anode thereby providing more or less the same conductance for all parts of cathode. Presence of colloidal matter improves the throwing power but increase of temperature may produce the opposite effect.

Extraction of Metals

This is done in two ways:

1. The ore is treated with a strong acid to obtain a salt and the solution of such a salt is electrolyzed to liberate the metal.
2. When the ore in molten state is available it is electolysed in a furnace.

Extraction of Zinc

The ore consisting of zinc is treated with concentrated sulphuric acid, roasted and passed through other processes to get rid of impurities by precipitation. The zinc-sulphate solution is then electrolysed. The cells consist of large lead-lined wooden boxes having aluminum cathodes and lead anodes. The current density is about 1000 amperes per square meter. Zinc is deposited on cathodes.

Extraction of Aluminium

Ores of aluminium are bauxite cryolite.Bauxite is treated chemically and reduced to aluminium oxide and then dissolved in fused cryolite and electrolysed.The furnce is lined with carbon.The temperature of the furnace is about 1000^0 C to keep the electrolyte in a fused state. Aluminium deposits at the cathode.

Refining of Metals

Electrolytic extraction gives about 98 to 99 percent pure metal. Further refining is done by electrolysis. The anodes are made of the impure metal extracted from its ores and the electrolyte is a solution of the salt of the metal. Pure metal is deposited on the cathode.

Example : 1

A 20 cm long portion of a circular shaft 10 cm diameter is to be coated with a layer of 1.5 mm nickel. Determine the quantity of electricity in Ah and the time taken for the process. Assume a current density of 195 A/sq.m and a current efficiency of 92 percent. Specific gravity of nickel is 8.9.

Solution :

Wt. of nickel = 8.9 gm/cm^3

Wt of nickel to be deposited

$$= \pi \times 10 \times \frac{1.5}{10} \times 8.9 \times 10^{-3} \text{ kg}$$

Electro-chemical equivalent of nickel is 1.0954 kg per 1,000Ah.

Quantity of electricity required

$$= \frac{838.4 \times 10^{-3} \times 1{,}000}{1.0954 \times 0.92} = 833 \text{ Ahr}$$

Current density =195 A/m^2 .

Time taken = $\frac{833}{\pi \times 10 \times 20 \times 10^{-4} \times 195}$ = 68 hours.

XXXXXXXXXX

CHAPTER-2
ELECTRICAL HEATING

Electric heating is extensively used both for domestic and industrial applications. Domestic applications include **(i)** room heaters **(ii)** immersion heaters for water heating **(iii)** hot plates for cooking **(iv)** electric kettles **(v)** electric irons **(vi)** pop-corn plants **(vii)** electric ovens for bakeries and **(viii)** electric toasters etc. Industrial applications of electric heating include **(i)** melting of metals **(ii)** heat treatment of metals like annealing, tempering, soldering and brazing etc. **(iii)** moulding of glass **(iv)**Baking of insulators **(v)** enamelling of copper wires etc.

Advantage of electrical heating:

As compared to other methods of heating using gas, coal and fire etc., electric heating is far superior for the following reasons:

(i) Cleanliness. Since neither dust nor ash is produced in electric heating, it is a clean system of heating requiring minimum cost of cleaning.

(ii) No Pollution. Since no flue gases are produced in electric heating, no provision has to be made for their exit.

(iii) Economical. Electric heating is economical because electric furnaces are cheaper in their initial cost as well as maintenance cost since they do not require big space for installation or for storage of coal and wood. Moreover, there is no need to construct any chimney or to provide extra heat installation.

(iv)Ease of Control. It is easy to control and regulate the temperature of an electric furnace with the help of manual or automatic devices. Temperature can be controlled within ± 5°C which is not possible in any other form of heating.

(v) Special Heating Requirement. Special heating requirements such as uniform heating of a material or heating one particular portion of the job without affecting its other parts or heating with no oxidation can be met only by electric heating.

(vi) Higher Efficiency. Heat produced electrically does not go away waste through the chimney and other by products. Consequently, most of the heat produced is utilised for heating the material itself. Hence, electric heating has higher efficiency as compared to other types of heating.

(vii) Better Working Conditions. Since electric heating produces no irritating noises and also the radiation losses are low, it results in low ambient temperature. Hence, working with electric furnaces is convenient and cool.

(viii) Heating of Bad Conductors. Bad conductors of heat and electricity like wood, plastic and bakery items can be uniformly and suitably heated with dielectric heating process.

(ix) Safety. Electric heating is quite safe because it responds quickly to the controlled signals.

(x) Lower Attention and Maintenance Cost. Electric heating equipment generally will not require much attention and supervision and their maintenance cost is almost negligible. Hence, labour charges are negligibly small as compared to other forms of heating.

Different Methods of Heat Transfer

The different methods by which heat is transferred from a hot body to a cold body are as under:

I. Conduction
II. Convection
III. Radiation

I. Conduction

In this mode of heat transfer, one molecule of the body gets heated and transfers some of the heat to the adjacent molecule and so on. There is a temperature gradient between the two ends of the body being heated.

Consider a solid material of cross-section Asq.m. and thickness x metre as shown in Fig.1.

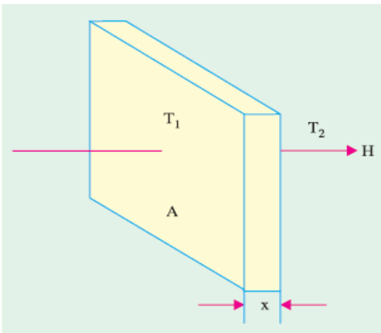


Fig-1

If T1 and T2 are the temperatures of the two sides of the slab in °K, then heat conducted between the two opposite faces in time t seconds is given by:

$$H = \frac{KA\,(T_1 - T_2)t}{x} \quad \ldots\ldots(1)$$

Where, K is thermal conductivity of the material.

II. Convection

In this process, heat is transferred by the flow of hot and cold air currents. This process is applied in the heating of water by immersion heater or heating of buildings. The quantity of heat absorbed by the body by convection process depends mainly on the temperature of the heating element above the surroundings and upon the size of the surface of the heater. It also depends, to some extent, on the position of the heater. The amount of heat dissipated is given by H = a (T1 – T2), where a is constant and T1 and T2 are the temperatures of the heating surface and the fluid in °K respectively. In electric furnaces, heat transferred by convection is negligible.

III. Radiation

It is the transfer of heat from a hot body to a cold body in a straight line without affecting the intervening medium. The rate of heat emission is given by Stefan's law, according to which heat dissipated is given by equation—2.

$$H = 5.72\, eK\left[\left(\frac{T_1}{100}\right)^4 - \left(\frac{T_2}{100}\right)^4\right] \quad W/m^2 \quad \dots\dots(2)$$

Where, K is radiating efficiency and e is known as emissivity of the heating element. If d is the diameter of the heating wire and $l$its total length, then its surface area from which heat is radiated,

S = πdl .... (3)

If H is the power radiated per m^2 of the heating surface, then,

Total power radiated as heat = $H\pi dl$....(4)

If P is the electrical power input to the heating element, then

$$P = \pi dl \times H \quad \dots..(5)$$

Resistance Heating.

It is based on the I^2R effect. When current is passed through a resistance element, I^2R loss takes place which produces heat. There are two methods of resistance heating.

(a) Direct Resistance Heating.

In this method the material (or charge) to be heated is treated as a resistance and current is passed through it. The charge may be in the form of powder, small solid pieces or liquid. The two electrodes are inserted in the charge and connected to either a.c. or d.c. supply (Fig. 2). Obviously, two electrodes will be required in the case of d.c. or single-phase a.c. supply but there would be three electrodes in the case of 3-phase supply. When the charge is in the form of small pieces, a powder of high resistivity material is sprinkled over the surface of the charge to avoid direct short circuit. Heat is produced when current passes through it. This method of heating has high efficiency because the heat is produced in the charge itself.

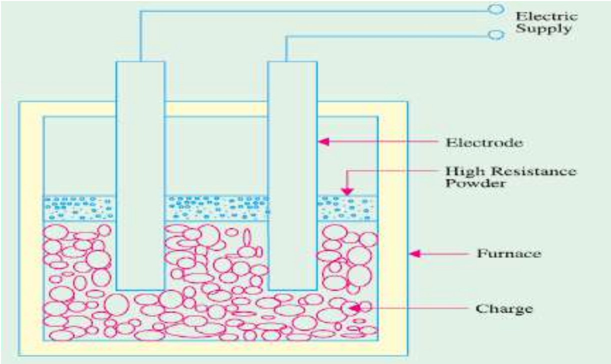


Fig:2 Direct Resistance heating

b) In-Direct Resistance heating.

In this method of heating, electric current is passed through a resistance element which is placed in an electric oven. Heat produced is proportional to I^2R losses in the heating element. The heat so produced is delivered to the charge either by radiation or convection or by a combination of the two. Sometimes, resistance is placed in a cylinder which is surrounded by the charge placed in the jacket as shown in the Fig.3. This arrangement provides uniform temperature. Moreover, automatic temperature control can also be provided.

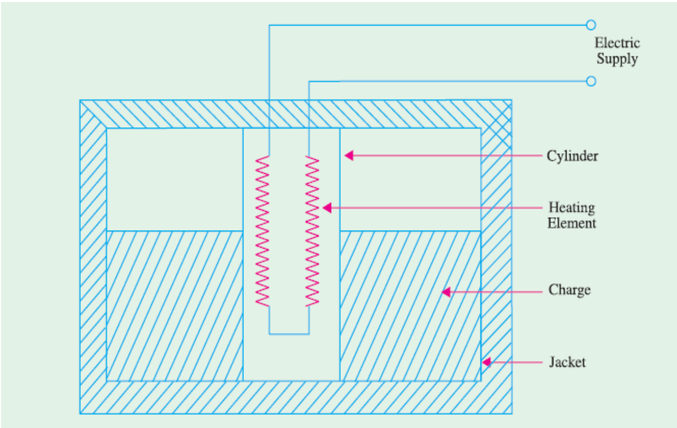


Fig-3 Indirect Resistance heating

<u>Principle of Resistance furnace.</u>

These are suitably-insulated closed chambers with a provision for ventilation and are used for a wide variety of purposes including heat treatment of metals like annealing and hardening etc., staving of enamelled wares, drying and baking of potteries, vulcanizing and hardening of synthetic materials and for commercial and domestic heating. Temperatures up to 1000°C can be obtained by using heating elements made of nickel, chromium and iron. Ovens using heating elements made of graphite can produce temperatures up to 3000°C.

Heating elements may consist of circular wires or rectangular ribbons. The ovens are usually made of a metal framework having an internal lining of fire bricks. The heating element may be located on the top, bottom or sides of the oven. The nature of the insulating material is determined by the maximum temperature required in the oven. An enclosure for charge which is heated by radiation or convection or both is called a **heating chamber**.

Electric resistance furnace

Fig. 4

Temperature Control of Resistance Furnaces

The temperature of a resistance furnace can be changed by controlling the I^2R or V^2/R losses.

Following different methods are used for the above purpose:

(1) Intermittent Switching.

In this case, the furnace voltage is switched ON and OFF intermittently. When the voltage supply is switched off, heat production within the surface is stalled and hence its temperature is reduced.

When the supply is restored, heat production starts and the furnace temperature begin to increase. Hence, by this simple method, the furnace temperature can be limited between two limits.

(2) By Changing the Number of Heating Elements.

In this case, the number of heating elements is changed without cutting off the supply to the entire furnace. Smaller the number of heating elements, lesser the heat produced. In the case of a 3-phase circuit, equal number of heating elements is switched off from each phase in order to maintain a balanced load condition.

(3) Variation in Circuit Configuration.

In the case of 3-phase secondary load, the heating elements give less heat when connected in a star than when connected in delta because in the two cases, voltages across the elements is different (Fig.5). In single-phase circuits, series and parallel grouping of the heating elements causes change in power dissipation resulting in change of furnace temperature. As shown in Fig.6 heat produced is more when all these elements are connected in parallel than when they are connected in series or series-parallel.

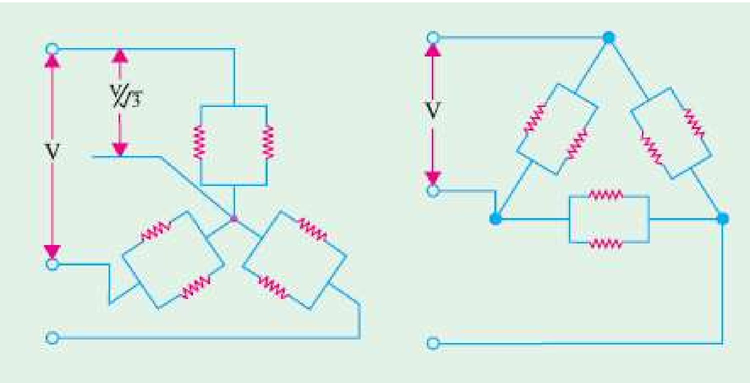


Fig-5

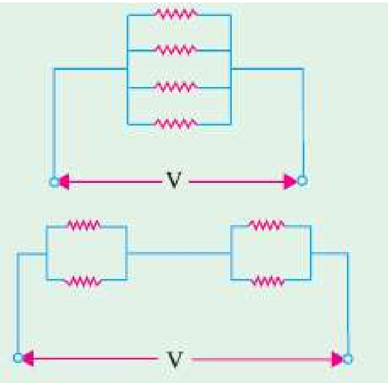


Fig-6

(4) Change of Applied Voltage.

(a) In the case of a furnace transformer having high voltage primary, the tapping control is kept in the primary winding because the magnitude of the primary current is less. Consider the multi-tap step-down transformer shown in Fig.7.

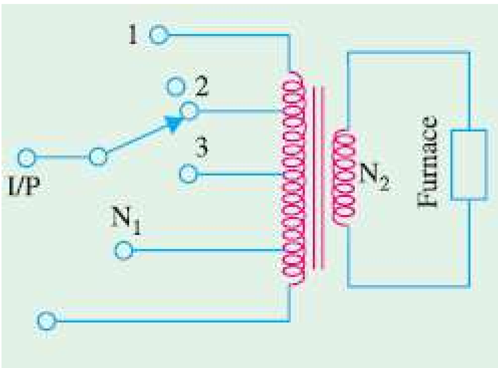


Fig-7

Let the four tapings on the primary winding have 100%, 80%, 60% and 50%. When 100% primary turns are used, secondary voltage is given by $V_2 = (N_2/N_1)V_i$, where Vi is the input voltage. When 50% tapping is used, the number of primary turns involved is N1/2. Hence, available secondary voltage $V_2 = (2N_2/N_1)Vi$. By selecting a suitable primary tapping, secondary voltage can be increased or decreased causing a change of temperature in the furnace.

(b) Bucking-Boosting the Secondary Voltage.

In this method, the transformer secondary is wound in two sections having unequal number of turns. If the two sections are connected in series aiding, the secondary voltage is boosted i.e., increased to (E2 + E3) as shown in Fig.8 (a). When the two sections are connected in series-opposing [Fig.8(b)] the secondary voltage is reduced i.e., there is bucking effect. Consequently, furnace voltage becomes (E2 – E3) and, hence, furnace temperature is reduced.

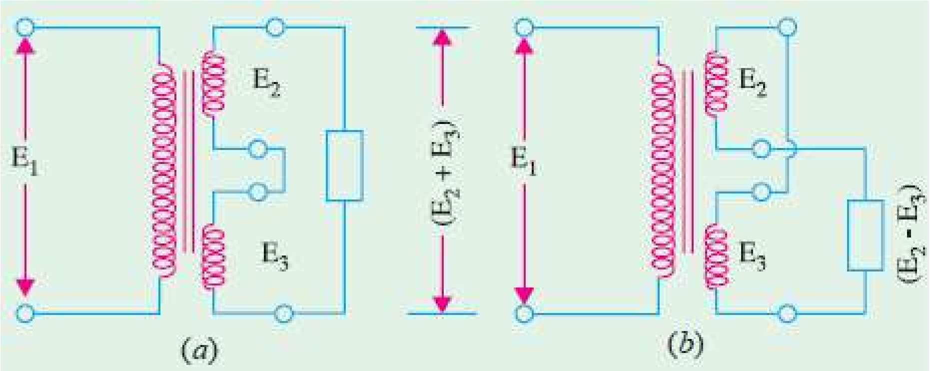


Fig-8

(c) Autotransformer Control.

Fig.9 shows the use of tapped autotransformer used for decreasing the furnace voltage and, hence, temperature of small electric furnaces. The required voltage can be selected with the help of a voltage selector.

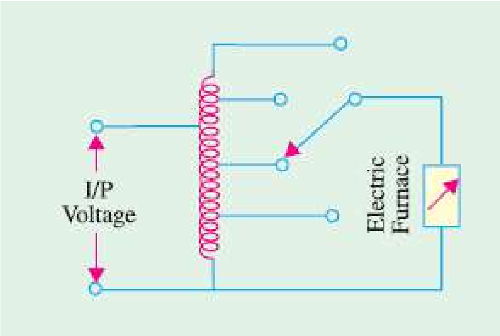


Fig-9

(d) Series Reactor Voltage.

In this case, a heavy-duty core-wounded coil is placed in series with the furnace as and when desired. Due to drop in voltage across the impedance of the coil, the voltage available across the furnace is reduced. With the help of D.P.D.T. switch, high/low, two mode temperature control can be obtained as shown in the Fig.10. Since the addition of series coil reduces the power factor, a power capacitor is simultaneously introduced in the circuit for keeping the p.f. nearly unity. As seen, the inductor is connected in series, whereas the capacitor is in parallel with the furnace.

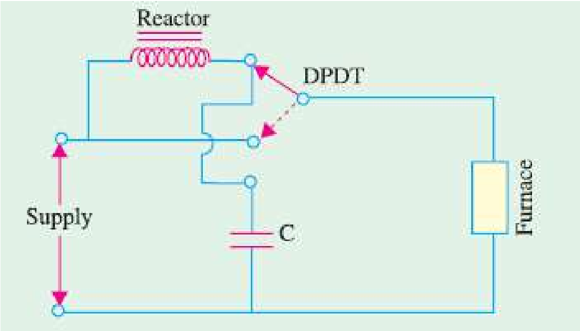


Fig-10

Arc Furnaces

If a sufficiently high voltage is applied across an air-gap, the air becomes ionized and starts conducting in the form of a continuous spark or arc thereby producing intense heat. When electrodes are made of carbon/graphite, the temperature obtained is in the range of 3000°C-3500C. The high voltage required for striking the arc can be obtained by using a step-up transformer fed from a variable a.c. supply as shown in Fig. 11 (a).

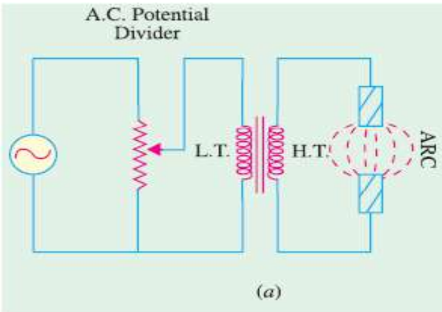


Fig-11

Indirect Arc Furnace

In this case, arc is formed between the two electrodes and the charge in such a way that electric current passes through the body of the charge as shown in Fig.11(a) . Such furnaces produce very high temperatures. In this case, arc is formed between the two electrodes and the heat thus produced is passed on to the charge by radiation as shown in Fig. 47.11 (b).

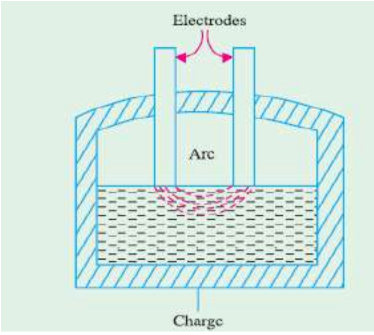


Fig-11(a)

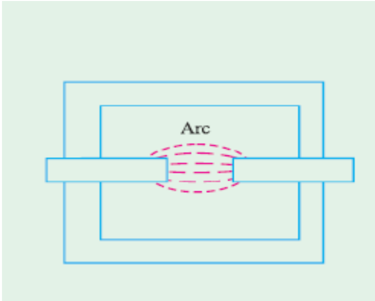


Fig-11(b)

Direct Arc Furnace

It could be either of conducting-bottom type [Fig.12 (a)] or non-conducting bottom type [Fig.12 (b)]. As seen from Fig.12 (a), bottom of the furnace forms part of the electric circuit so that current passes through the body of the charge which offers very low resistance. Hence, it is possible to obtain high temperatures in such furnaces. Moreover, it produces uniform heating of charge without stirring it mechanically. In Fig.12 (b), no current passes through the body of the furnace. Most common application of these furnaces is in the production of steel because of the ease with which the composition of the final product can be controlled during refining. Most of the furnaces in general use are of non-conducting bottom type due to insulation problem faced in case of conducting bottom.

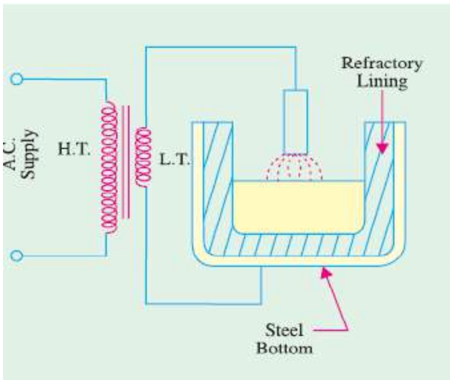


Fig-12(a)

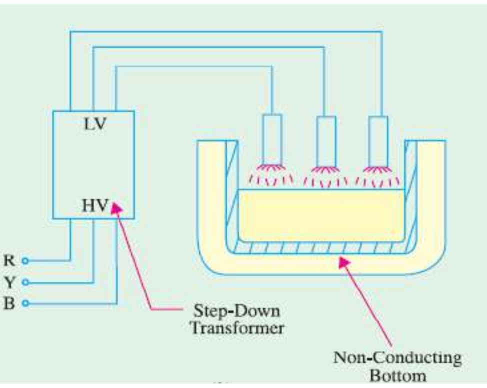


Fig-12(b)

Indirect Arc Furnace

Fig.13 shows a single-phase indirect arc furnace which is cylindrical in shape. The arc is struck by short circuiting the electrodes manually or automatically for a moment and then, withdrawing them apart. The heat from the arc and the hot refractory lining is transferred to the top layer of the charge by radiation. The heat from the hot top layer of the charge is further transferred to other parts of the charge by conduction. Since no current passes through the body of the charge, there is no inherent stirring action due to electro-magnetic forces set up by the current. Hence, such furnaces have to be rocked continuously in order to distribute heat uniformly by exposing different layers of the charge to the heat of the arc. An electric motor is used to operate suitable grinders and rollers to impart rocking motion to the furnace. Rocking action provides not only thorough mixing of the charge, it also increases the furnace efficiency in addition to increasing the life of the refractory lining material. Since in this furnace, charge is heated by radiation only, its temperature is lower than that obtainable in a direct arc furnace. Such furnaces are mainly used for melting nonferrous metals although they can be used in iron foundries where small quantities of iron are required frequently.

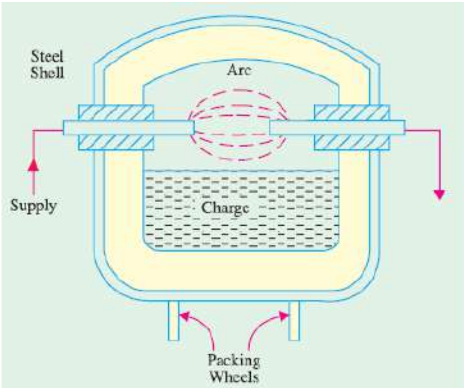


Fig-13

Induction Heating

This heating process makes use of the currents induced by the electro-magnetic action in the charge to be heated. In fact, induction heating is based on the principle of transformer working. The primary winding which is supplied from an a.c. source is magnetically coupled to the charge which acts as a short circuited secondary of single turn. When an a.c. voltage is applied to the primary, it induces voltage in the secondary i.e. charge. The secondary current heats up the charge in the same way, as any electric current does while passing through a resistance. If V is the voltage induced in the charge and R is the charge resistance, then heat produced = V 2/R. The value of current induced in the charge depends on **(i)** magnitude of the primary current **(ii)** turn ratio of the transformer **(iii)** co-efficient of magnetic coupling. Low-frequency induction furnaces are used for melting and refining of different metals. However, for other processes like case hardening and soldering etc., high frequency eddy-current heating is employed. Low frequency induction furnaces employed for the melting of metals are of the following two types:

(a) Core-type Furnaces — It operates just like a two winding transformer. These can be further sub-divided into **(i)** Direct core-type furnaces **(ii)** Vertical core-type furnaces and **(iii)** Indirect core-type furnaces.

(b) Coreless-type Furnaces — in which an inductively-heated element is made to transfer heat to the charge by radiation.

<u>Core Type Induction Furnace</u>

It is shown in Fig.14 and is essentially a transformer in which the charge to be heated forms a single-turn short-circuited secondary and is magnetically coupled to the primary by an iron core. The furnace consists of a circular hearth which contains the charge to be melted in the form of an annular ring. When there is no molten metal in the ring, the secondary becomes open-circuited there-by cutting off the secondary current. Hence, to start the furnace, molten metal has to be poured in the annular hearth. Since, magnetic coupling between the primary and secondary is very poor, it results in high leakage and low power factor. In order to nullify the effect of increased leakage reactance, low primary frequency of the order of 10 Hz is used. If the transformer secondary current density exceeds 500 A/cm2 then, due to the interaction of secondary current with the alternating magnetic field, the molten metal is squeezed to the extent that secondary circuit is interrupted. This effect is known as "pinch effect".

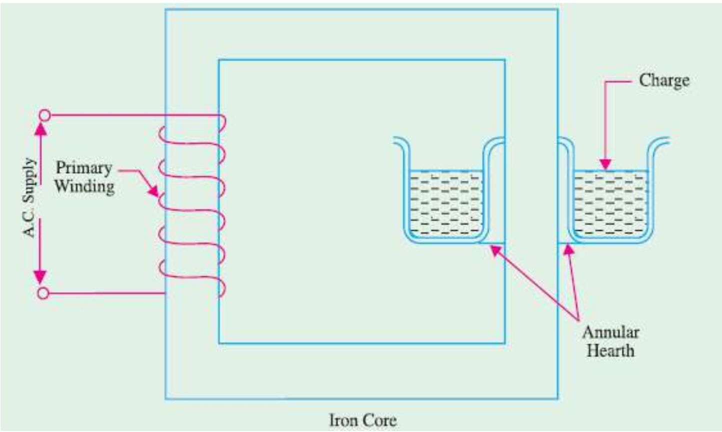


Fig-14

This furnace suffers from the following drawbacks:

1. It has to be run on low-frequency supply which entails extra expenditure on motor-generator set or frequency convertor.
2. It suffers from pinching effect.
3. The crucible for charge is of odd shape and is very inconvenient for tapping the molten charge.
4. It does not function if there is no molten metal in the hearth i.e. when the secondary is open. Every time molten metal has to be poured to start the furnace.
5. It is not suitable for intermittent service. However, in this furnace, melting is rapid and clean and temperature can be controlled easily. Moreover, inherent stirring action of the charge by electro-magnetic forces ensures greater uniformity of the end product.

Vertical Core-Type Induction Furnace

It is also known as Ajax-Wyatt furnace and represents an improvement over the core-type furnace discussed above. As shown in Fig.15, it has vertical channel (instead of a horizontal one) for the charge, so that the crucible used is also vertical which is convenient from metallurgical point of view. In this furnace, magnetic coupling is comparatively better and power factor is high. Hence, it can be operated from normal frequency supply. The circulation of the molten metal is kept up round the Vee portion by convection currents as shown in Fig.15. As Vee channel is narrow, even a small quantity of charge is sufficient to keep the secondary circuit closed. However, Vee channel must be kept full of charge in order to maintain continuity of secondary circuit. This fact makes this furnace suitable for continuous operation. The tendency of the secondary circuit to rupture due to pinch-effect is counteracted by the weight of the charge in the crucible. The choice of material for inner lining of the furnace depends on the type of charge used. Clay lining is used for yellow brass. For red brass and bronze, an alloy of magnetia and alumina or corundum is used. The top of the furnace is covered with an insulated cover which can be removed for charging. The furnace can be tilted by the suitable hydraulic arrangement for taking out the molten metal. This furnace is widely used for melting and refining of brass and other non-ferrous metals. As said earlier, it is suitable for continuous operation. It has a p.f. of 0.8-0.85. With normal supply frequency, its efficiency is about 75% and its standard size varies from 60-300 kW, all single phase.

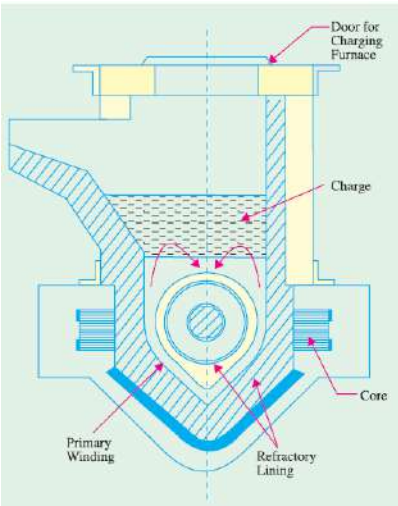


Fig-15 Core type Induction furnace

Indirect Core-Type Induction Furnace

In this furnace, a suitable element is heated by induction which, in turn, transfers the heat to the charge by radiation. So far as the charge is concerned, the conditions are similar to those in a resistance oven. As shown in Fig.16, the secondary consists of a metal container which forms the walls of the furnace proper. The primary winding is magnetically coupled to this secondary by an iron core. When primary winding is connected to a.c. supply, secondary current is induced in the metal container by transformer action which heats up the container.

The metal container transfers this heat to the charge. A special advantage of this furnace is that its temperature can be automatically controlled without the use of an external equipment. The part AB of the magnetic circuit situated inside the oven chamber consists of a special alloy which loses its magnetic properties at a particular temperature but regains them when cooled back to the same temperature. As soon as the chamber attains the critical temperature, reluctance of the magnetic circuit increases manifold thereby cutting off the heat supply. The bar AB is detachable and can be replaced by other bars having different critical temperatures.

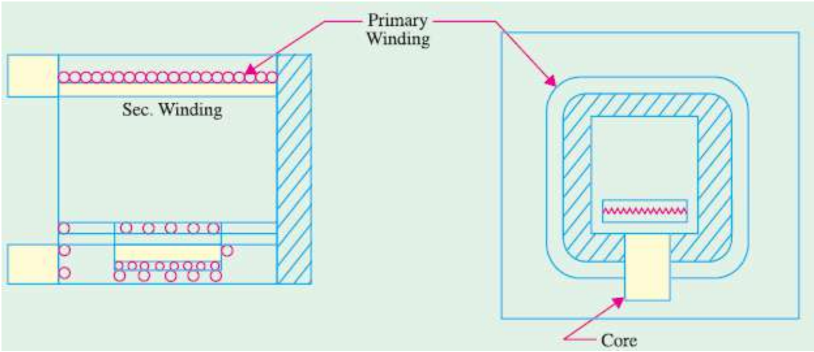


Fig-16

Coreless Induction Furnace

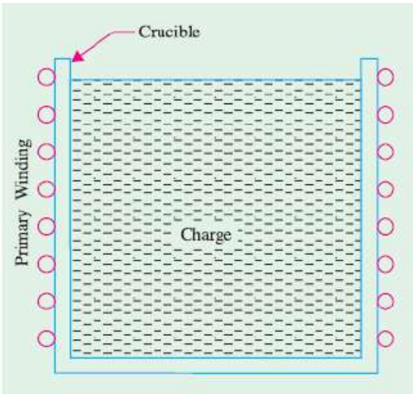


Fig-17

As shown in Fig.17, the three main parts of the furnace are **(i)** primary coil **(ii)** a ceramic crucible containing charge which forms the secondary and **(iii)** the frame which includes supports and tilting mechanism. The distinctive feature of this furnace is that it contains no heavy iron core with the result that there is no continuous path for the magnetic flux. The crucible and the coil are relatively light in construction and can be conveniently tilted for pouring. The charge is put into the crucible and primary winding is connected to a high-frequency a.c. supply. The flux produce by the primary sets up eddy-currents in the charge and heats it up to the melting point. The charge need not be in the molten state at the start as was required by core-type furnaces. The eddy- currents also set up electromotive forces which produce stirring action which is essential for obtaining uniforms quality of metal. Since flux density is low (due to the absence of the magnetic core) high frequency supply has to be used because eddy-current loss $We \propto B^2_{max} f^{\ 2}$. However, this high frequency increases the resistance of the primary winding due to skin effect, thereby increasing primary Cu losses. Hence, the primary winding is not made of Cu wire but consists of hollow Cu tubes which are cooled by water circulating through them. Since magnetic coupling between the primary and secondary windings is low, the furnace p.f. lies between 0.1 and 0.3. Hence, static capacitors are invariably used in parallel with the furnace to improve its p.f. Such furnaces are commonly used for steel production and for melting of non-ferrous metals like brass, bronze, copper and aluminium etc., along with various alloys of these elements. Special application of these furnaces include vacuum melting, melting in a controlled atmosphere and melting for precision casting where high frequency induction heating is used. It also finds wide use in electronic industry and in other industrial activities like soldering, brazing hardening and annealing and sterilizing surgical instruments etc. Some of the advantages of coreless induction furnaces are as follows:

1. They are fast in operation.
2. They produce most uniform quality of product.
3. They can be operated intermittently.
4. Their operation is free from smoke, dirt, dust and noises.
5. They can be used for all industrial applications requiring heating and melting.
6. They have low erection and operating costs.
7. Their charging and pouring is simple.

Dielectric Heating

It is also called high-frequency capacitive heating and is used for heating insulators like wood, plastics and ceramics etc. which cannot be heated easily and uniformly by other methods. The supply frequency required for dielectric heating is between 10-50 MHz and the applied voltage is up to 20 kV. The overall efficiency of dielectric heating is about 50%.

Dielectric Loss

When a practical capacitor is connected across an a.c. supply, it draws a current which leads the voltage by an angle φ, which is a little less than 90° or falls short of 90° by an angle δ. It means that there is a certain component of the current which is in phase with the voltage and hence produces some loss called dielectric loss. At the normal supply frequency of 50 Hz, this loss is negligibly small but at higher frequencies of 50 MHz or so, this loss becomes so large that it is sufficient to heat the dielectric in which it takes place. The insulating material to be heated is placed between two conducting plates in order to form a parallel-plate capacitor as shown in Fig.19 (a). Fig.19 (b) shows the equivalent circuit of the capacitor and Fig.19 (c) gives its vector diagram.

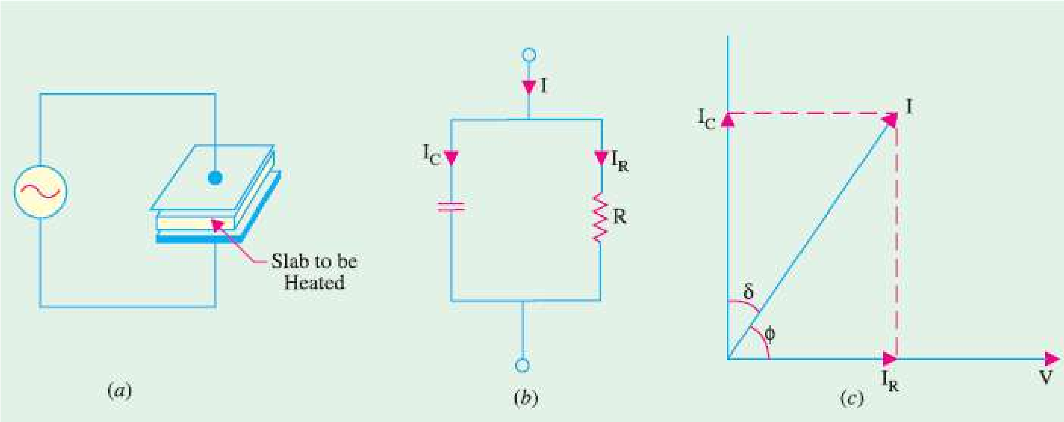


Fig-19

Power drawn from supply = VI cos φ

Now, Ic= I =V/Xc= 2πf CV

∴P = V(2πfCV) cos φ = 2πfCV^2 cos φ

Now, φ = (90^0 – δ), cos φ = cos (90^0 –δ) = sin δ = tan δ = δ

whereδ is very small and is expressed in radians.

P = 2πfCV^{2}δ watts

Here,

$$C = \epsilon_0 \epsilon_r \frac{A}{d}$$

Where, d is the thickness and A is the surface area of the dielectric slab.

This power is converted into heat. Since for a given insulator material, C and δ are constant, the dielectric loss is directly proportional to V 2 f. That is why high-frequency voltage is used in dielectric heating. Generally, a.c. voltage of about 20 kV at a frequency of 10-30 MHz is used.

<u>Advantages of Dielectric Heating</u>

1. Since heat is generated within the dielectric medium itself, it results in uniform heating.
2. Heating becomes faster with increasing frequency.
3. It is the only method for heating bad conductors of heat.
4. Heating is fastest in this method of heating.
5. Since no naked flame appears in the process, inflammable articles like plastics and wooden products etc. can be heated safely.
6. Heating can be stopped immediately as and when desired.

CHAPTER-3
WELDING

Definition

It is the process of joining two pieces of metal or non-metal at faces rendered plastic or liquid by the application of heat or pressure or both. Filler material may be used to effect the union.

Welding Processes

All welding processes fall into two distinct categories:

1. Fusion Welding—it involves melting of the parent metal. Examples are:

(i) Carbon arc welding, metal arc welding, electron beam welding, electro-slag welding and electro-gas welding which utilize electric energy and

(ii) Gas welding and thermal welding which utilize chemical energy for the melting purpose.

2. Non-fusion Welding—It does not involve melting of the parent metal. Examples are:

(i) Forge welding and gas non-fusion welding which use chemical energy.

(ii) Explosive welding, friction welding and ultrasonic welding etc., which use mechanical energy.

(iii) Resistance welding which uses electrical energy.

Proper selection of the welding process depends on the **(a)** kind of metals to be joined **(b)** cost involved **(c)** nature of products to be fabricated and **(d)** production techniques adopted.

Use of Electricity in Welding

Electricity is used in welding for generating heat at the point of welding in order to melt the material which will subsequently fuse and form the actual weld joint. There are many ways of producing this localised heat but the two most common methods are as follows:

1. Resistance welding—here current is passed through the inherent resistance of the joint to be welded thereby generating the heat as per the equation I^2Rt/J kilocalories.

2. Arc welding—here electricity is conducted in the form of an arc which is established between the two metallic surfaces

Principle of arc welding

Formation and Characteristics of Electric Arc:

An electric arc is formed whenever electric current is passed between two metallic electrodes which are separated by a short distance from each other. The arc is started by momentarily touching the positive electrode (anode) to the negative metal (or plate) and then withdrawing it to about 3 to 6 mm from the plate. When electrode first touches the plate, a large short-circuits current flows and as it is later withdrawn from the plate, current continues to flow in the form of a spark across the air gap so formed. Due to this spark (or discharge), the air in the gap becomes ionized i.e. is split into negative electrons and positive ions. Consequently, air becomes conducting and current is able to flow across the gap in the form of an arc. As shown in Fig. 48.2, the arc consists of **lighter** electrons which flow from cathode to anode and **heavier** positive ions which flow from anode to cathode. Intense heat is generated when high velocity electrons strike the anode. Heat generated at the cathode is much less because of the low velocity of the impinging ions. It is found that nearly **two-third** of the heat is developed at the anode which burns into the form of a crater where temperature rises to a value of 3500-4000°C. The remaining one-third of the heat is developed near the cathode. The above statement is true in all d.c. systems of welding where positive side of the circuit is the hottest side. As a result, an electrode connected to the positive end of the d.c. supply circuit will burn 50% faster than if connected to the negative end. This fact can be used for obtaining desired penetration of the base metal during welding.

Four Positions of Arc Welding

There are four basic positions in which manual arc welding is done.

1. Flat position. It is shown in Fig.20 (a). Of all the positions, flat position is the easiest, most economical and the most used for all shielded arc welding. It provides the strongest weld joints. Weld beads are exceedingly smooth and free of slag spots. This position is most adaptable for welding of both ferrous and non-ferrous metals particularly for cast iron.

2. Horizontal Position. It is the second most popular position and is shown in Fig.20(b). It also requires a short arc length because it helps in preventing the molten puddle of the metal from sagging. However, major errors that occur while welding in horizontal position are under-cutting and over-lapping of the weld zone .

3. Vertical Position. It is shown in Fig.20(c). In this case, the welder can deposit the bead either in the uphill or downhill direction. Downhill welding is preferred for thin metals because it is faster than the uphill welding. Uphill welding is suited for thick metals because it produces stronger welds.

4. Overhead Position. It is shown in Fig.20(d). Here, the welder has to be very cautious otherwise he may get burnt by drops of falling metal. This position is thought to be the most hazardous but not the most difficult one.

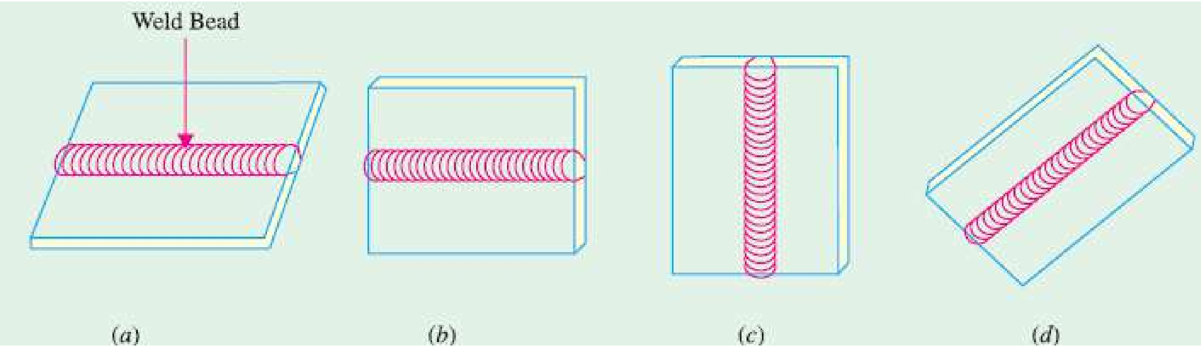


Fig-1

Electrodes for Metal Arc Welding

An electrode is a filler metal in the form of a wire or rod which is either bare or coated uniformly with flux. As per IS : 814-1970, the contact end of the electrode is left bare and clean to a length of 20-30 mm. for inserting it into electrode holder (Fig.21).

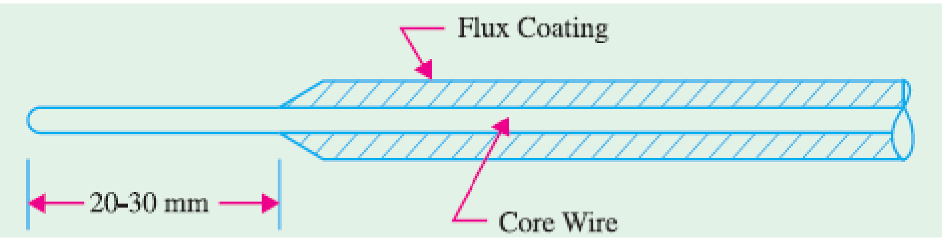


Fig-2

Metal arc welding was originally done with bare electrodes which consisted of a piece of wire or rod of the same metal as the base metal. However, due to atmospheric contamination, they produced brittle and poor quality welds. Hence, bare wire is no longer used except for automatic welding in which case arrangement is made to protect the weld area from the atmosphere by either powdered flux or an inert gas. Since 1929, coated electrodes are being extensively used for shielded arc welding. They consist of a metal core wire surrounded by a thick flux coating applied by extrusion, winding or other processes. Depending on the thickness of the flux coating, coated electrodes may be classified into **(i)** lightly-dusted (or dipped) electrodes and **(ii)** semi-coated (or heavy coated) electrodes. Materials commonly used for coating are **(i)** titanium oxide **(ii)** ferromanganese **(iii)** silica flour **(iv)** asbestos clay **(v)** calcium carbonate and **(vi)** cellulose with sodium silicate often used to hold ingredients together. Electrode coating contributes a lot towards improving the quality of the weld. Part of the coating burns in the intense heat of the arc and provides a gaseous shield around the arc which prevents oxygen, nitrogen and other impurities in the atmosphere from combining with the molten metal to cause a poor quality brittle and weak weld. Another portion of the coating flux melts and mixes with the impurities in the molten pool causing them to float to the top of the weld where they cool in the form of slag. This slag improves the bead quality by protecting it from the contaminating effects of the atmosphere and causing it to cool down more uniformly. It also helps in controlling the basic shape of the weld bead. The type of electrode used depends on the type of metal to be welded, the welding position, the type of electric supply whether a.c. or d.c. and the polarity of the welding machine.

Carbon Arc Welding

(a) General

Carbon arc welding was the first electric welding process developed by a French inventor Auguste de Meritens in 1881. In this process, fusion of metal is accomplished by the heat of an electric arc. No pressure is used and generally, no shielding atmosphere is utilized. Filler rod is used only when necessary. Although not used extensively these days, it has, nevertheless, certain useful fields of application. Carbon arc welding differs from the more common shield metal arc welding in that **it uses non-consumablecarbon or graphic electrodes** instead of the consumable flux-coated electrodes.

(b) Welding Circuit

The basic circuit is shown in Fig.22 and can be used with d.c. as well as a.c. supply. When direct current is used, the electrode is mostly negative (DCSP). The process is started by adjusting the amperage on the d.c.welder, turning welder ON and bringing the electrode into contact with the work piece. After the arc column starts, electrode is withdrawn 25 – 40 mm away and the arc is maintained at this distance. The arc can be extinguished by simply removing the electrode from the work piece completely. The only function of the carbon arc is to supply heat to the base metal. This heat is used to melt the base metal or filler rod for obtaining fusion weld Depending on the type and size of electrodes, maximum current values range from 15 A to 600 A for single-electrode carbon arc welding.

(c) Electrodes

These are made of either carbon or graphite, are usually 300 mm long and 2.5 – 12 mm in diameter. Graphite electrodes are harder, more brittle and last longer than carbon electrodes. They can withstand higher current densities but their arc column is harder to control. Though considered non-consumable, they do disintegrate gradually due to vaporisation and oxidisation.

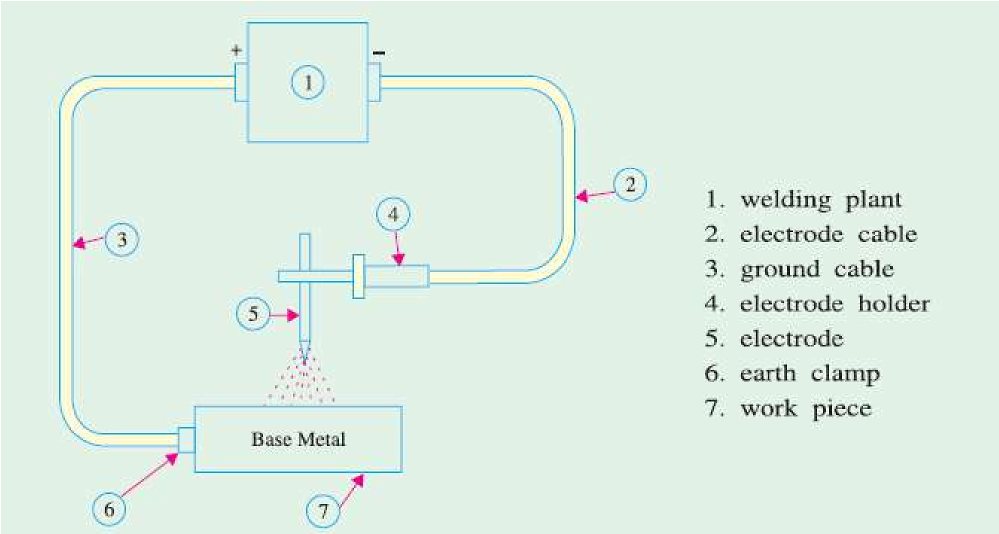


Figure. 3

(d) Applications

1. The joint designs that can be used with carbon arc welding are butt joints, bevel joints, flange joints, lap joints and fillet joints.

2. This process is easily adaptable for automation particularly where amount of weld deposit is large and materials to be fabricated are of simple geometrical shapes such as water tanks.

3. It is suitable for welding galvanised sheets using copper-silicon-manganese alloy filler metal.

4. It is useful for welding thin high-nickel alloys.

5. Monel metal can be easily welded with this process by using a suitable coated filler rod.

6. Stainless steel of thinner gauges is often welded by the carbon-arc process with excellent results.

(e) Advantages and Disadvantages

1. The main advantage of this process is that the temperature of the molten pool can be easily controlled by simply varying the arc length.
2. It is easily adaptable to automation.
3. It can be easily adapted to inert gas shielding of the weld and
4. It can be used as an excellent heat source for brazing, braze welding and soldering etc.

Its disadvantages are as under:

1. A separate filler rod has to be used if any filler material is required.
2. Since arc serves only as a heat source, it does not transfer any metal to help reinforce the weld joint.

3. The major disadvantage of the carbon-arc process is that blow holes occur due to magnetic arc blow especially when welding near edges of the work piece.

Submerged Arc Welding

In this **fusion** process, welding is done under a blanket of granulated flux which shields the weld from all bad effects of atmospheric gases while a consumable electrode is continuously and mechanically fed into the arc. The arc, the end of the bare metal electrode and the molten weld pool are all submerged under a thick mound of finely-divided granulated powder that contains deoxidisers, cleansers and other fluxing agents. The fluxing powder is fed from a hopper that is carried on the welding head itself (Fig.23). This hopper spread the powder in a continuous mound ahead of the electrode in the direction of welding. Since arc column is completely submerged under the powder, there is no splatter or smoke and, at the same time, weld is completely protected from atmospheric contamination. Because of this protection, weld beads are extremely smooth. The flux adjacent to the arc column melts and floats to the top of the molten pool where it solidifies to form slag. This slag is easy to remove. Often it cracks off by itself as it cools. The unused flux is removed and is reused again and again.

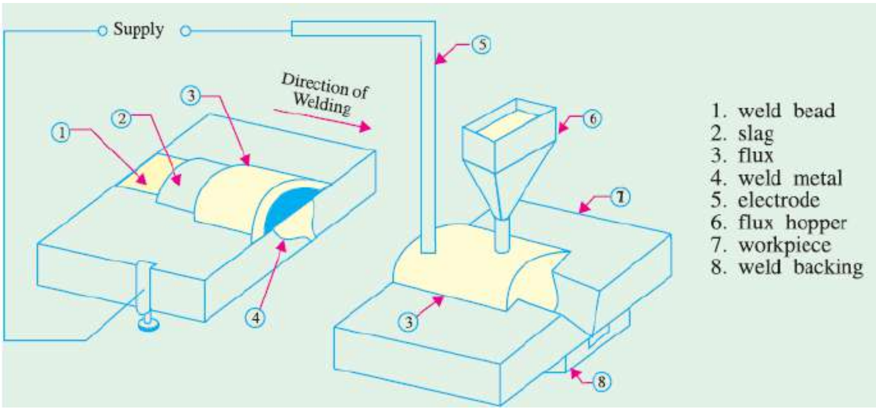


Fig-4

The electrode is either a bare wire or has a slight mist of copper coated over it to prevent oxidation. In automatic or semi-automatic submerged arc welding, wire electrode is fed mechanically through an electrically contacting collect. Though a.c. power supply may be used, yet d.c. supply is more popular because it assures a simplified and positive control of the welding process. This process requires high current densities about 5 to 6 times of those used in ordinary manual stick electrode welding. As a result, melting rate of the electrode as well as welding speed become much higher. Faster welding speed minimizes distortion and war page. The submerged arc process is suitable for

1. Welding low-alloy, high-tensile steels.

2. Welding mild, low-carbon steels.

3. Joining medium-carbon steel, heat-resistant steels and corrosion-resistant steels etc.

4. Welding nickel, and other non-ferrous metals like copper. This process has many industrial applications such as fabrication of pipes, boiler pressure vessels, railroad tank cars, structural shapes etc. which demand welding in a straight line. Welds made by this process have high strength and ductility. A major advantage of this process is that fairly thick sections can be welded in a single pass without edge preparation. Submerged arc welding can be done manually where automatic process is not possible such as on curved lines and irregular joints. Such a welding gun is shown in Fig-24. Both manual and automatic submerged arc processes are most suited for flat and slightly downhill welding positions.

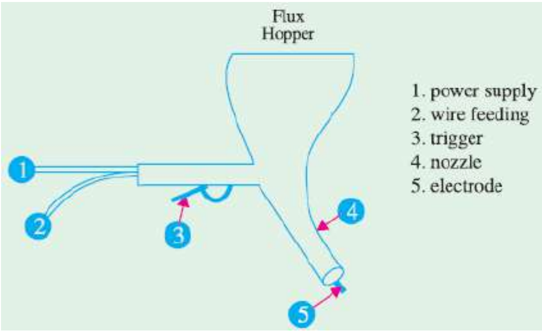


Fig-5

Twin Submerged Arc Welding

As shown in Fig.25, in this case, two electrodes are used simultaneously instead of one. Hence, weld deposit size is increased considerably. Moreover, due to increase in welding current (upto 1500 A), much deeper penetration of base metal is achieved.

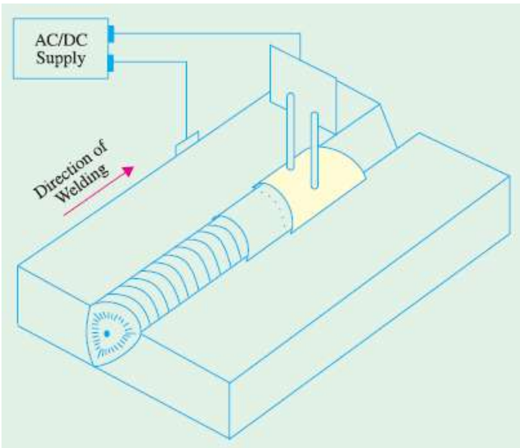


Fig-6

Gas Shield Arc Welding

In this fusion process, welding is done with bare electrodes but weld zone is shielded from the atmosphere by a gas which is piped to the arc column. Shielding gases used are carbon dioxide, argon, helium, hydrogen and oxygen. No flux is required. Different processes using shielding gas are as follows.

(a) Tungsten inert-gas (TIG) Process

In this process, non-consumable tungsten electrode is used and filler wire is fed separately. The weld zone is shielded from the atmosphere by the inert gas (argon or helium) which is ducted directly to the weld zone where it surrounds the tungsten and the arc column.

(b) Metal inert-gas (MIG) Process

It is a refinement of the TIG process. It uses a bare consumable (i.e. fusible) wire electrode which acts as the source for the arc column as well as the supply for the filler material. The weld zone is shielded by argon gas which is ducted directly to the electrode point.

Resistance Welding

It is fundamentally a heat and squeeze process. The term **'resistance welding'** denotes a group of processes in which welding heat is produced by the resistance offered to the passage of electric current through the two metal pieces being welded. These processes differ from the fusion processes in the sense that no extra metal is added to the joint by means of a filler wire or electrode. According to Joule's law, heat produced electrically is given by H = I2Rt/J. Obviously, amount of heat produced depends on. **(i)** square of the current **(ii)** the time of current and **(iii)** the resistance offered. As seen, in simple resistance welding, high-amperage current is necessary for adequate weld. Usually, R is the contact resistance between the two metals being welded together. The current is

passed for a suitable length of time controlled by a timer. The various types of resistance welding processes may be divided into the following four main groups :

(i) spot welding **(ii)** seam welding **(iii)** projection welding and **(iv)** butt welding which could be further subdivided into flash welding, upset welding and stud welding etc.

Advantages

Some of the advantages of resistance welding are as under :

1. Heat is localized where required
2. Welding action is rapid
3. No filler material is needed
4. Requires comparatively lesser skill
5. Is suitable for large quantity production
6. Both similar and dissimilar metals can be welded
7. Parent metal is not harmed

8. Difficult shapes and sections can be welded.

Only disadvantages are with regard to high initial as well as maintenance cost. It is a form of resistance welding in which the two surfaces are joined by spots of fused metal caused by fused metal between suitable electrodes under pressure.

Spot Welding

The process depends on two factors:

1. Resistance heating of small portions of the two work pieces to plastic state and

2. Application of forging pressure for welding the two work pieces. Heat produced is H = I2 Rt/J. The resistance R is made up of **(i)** resistance of the electrodes and metals themselves **(ii)** contact resistance between electrodes and work pieces and **(iii)** contact resistance between the two work pieces. Generally, contact resistance between the two work pieces is the greatest.

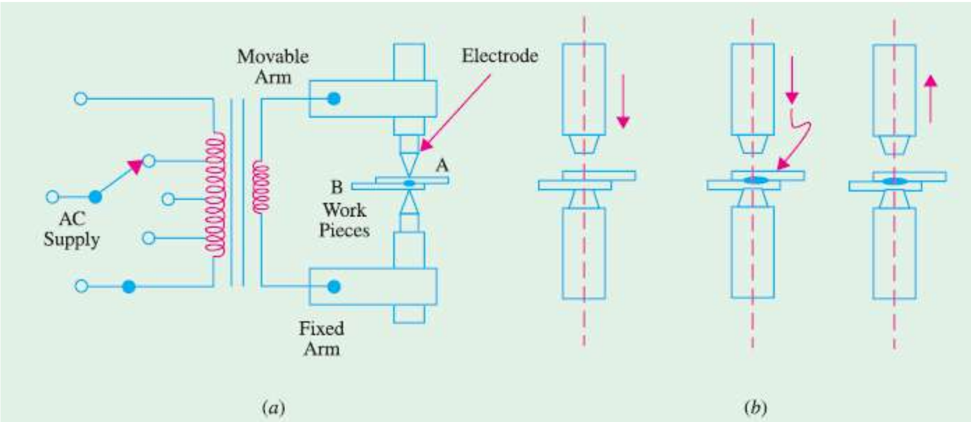


Fig-7

As shown in Fig-26 (b), mechanical pressure is applied by the tips of the two electrodes. In fact, these electrodes not only provide the forging pressure but also carry the welding current and concentrate the welding heat on the weld spot directly below them. Fig.26 (a) shows diagrammatically the basic parts of a modern spot welding. It consists of a step-down transformer which can supply huge currents (up to 5,000 A) for short duration of time. The lower arm is fixed whereas the upper one is movable. The electrodes are made of low-resistance, hard copper alloy and are either air cooled or butt-cooled by water circulating through the rifled drillings in the electrode. Pointed electrodes [Fig.27 (a)] are used for ferrous materials whereas domed electrodes are used for non-ferrous materials. Flat domes are used when spot-welding deformation is not desired. The weld size is determined by the diameter of the electrode.

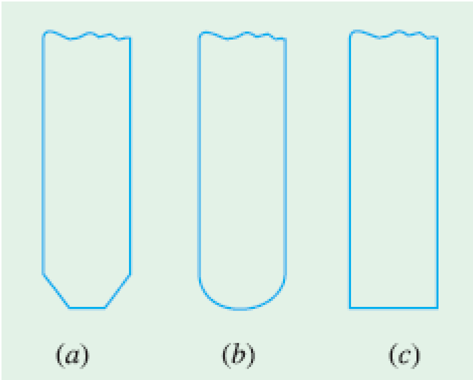


Fig-8

The welding machine is cycled in order to produce the required heat timed to coincide with the pressure exerted by the electrodes as shown in Fig.26 (a). As the movable electrode comes down and presses the two work pieces A and B together, current is passed through the assembly. The metals under the pressure zone get heated up to about 950°C and fuse together. As they fuse, their resistance is reduced to zero, hence there is a surge of current. This surge is made to switch off the welding current automatically. In motor-driven machines, speeds of 300strokes/minute are common. Spot welders are of two different types. One is a station arc welder which is available in different sizes. The other has a stationary transformer but the electrodes are in a gun form. Electric resistance spot welding is probably the best known and most widely-used because of its low cost, speed and dependability. It can be easily performed by even a semi-skilled operator. This process has a fast welding rate and quick set-up time apart from having low unit cost per weld. Spot welding is used for galvanized, tinned and lead coated sheets and mild steel sheet work. This technique is also applied to non-ferrous materials such as brass, aluminium, nickel and bronze etc.

Seam Welding

The seam welder differs from ordinary spot welder only in respect of its electrodes which are of disc or roller shape as shown in Fig.28(a). These copper wheels are power driven and rotate whilst gripping the work. The current is so applied through the wheels that the weld spots either overlap as in Fig.28 (b) or are made at regular intervals as in Fig.28 (c). The continuous or overlapped seam weld is also called **stitch weld** whereas the other is called roll weld.

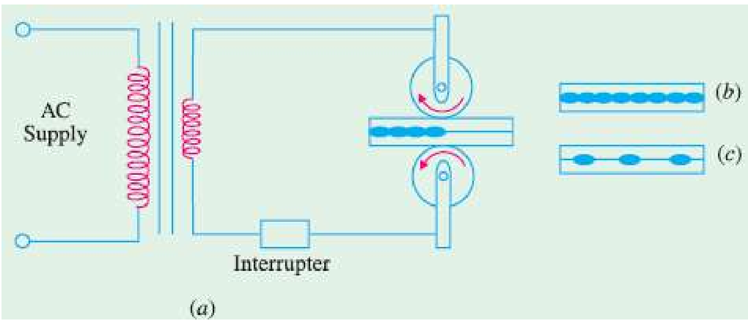


Fig-9

Seam welding is confined to welding of thin materials ranging in thickness from 2 mm to 5 mm. It is also restricted to metals having low harden ability rating such as hot-rolled grades of low alloy steels. Stitch welding is commonly used for long water-tight and gas-tight joints. Roll welding is used for simple joints which are not water-tight or gas-tight. Seam welds are usually tested by pillow test.

Projection Welding

It can be regarded as a mass-production form of spot welding. Technically, it is a cross between spot welding and butt welding. It uses the same equipment as spot welding. However, in this process, large-diameter flat electrodes (also called platens) are used. This welding process derives its name from the fact that, prior to welding, projections are raised on the surfaces to be welded [Fig.29 (a)].

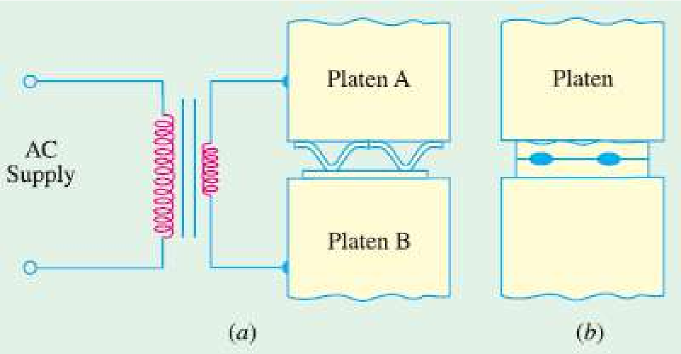


Fig-10

As seen, the upper and lower platens are connected across the secondary of a step-down transformer and are large enough to cover all the projections to be welded at one stroke of the machine. When platen A touches the work piece, welding current flows **through each projection**. The welding process is started by first lowering the upper platen A on to the work-piece and then applying mechanical pressure to ensure correctly-forged welds. Soon after, welding current is switched on as in spot welding. As projection areas heat up, they collapse and union takes place at all projections simultaneously [Fig.29(b)]. Projection welding is used extensively by auto manufactures for joining nuts, bolts and studs to steel plates in car bodies. This process is especially suitable for metals like brass, aluminium and copper etc. mainly due to their high thermal conductivity. A variation of projection welding is the metal fibre welding which uses a metal fibre rather than a projection point(Fig.30).

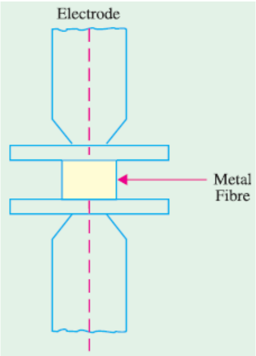


Fig-11

This metal fibre is generally a fill material. Instead of projections, tiny elements of this felt material are placed between the two metals which are then projection-welded in the usual way.

Butt Welding

In this case, the two work pieces are brought into contact end-to-end and the butted ends are heated by passing a heavy current through the joint. As in other forms of resistance welding, the weld heat is produced mainly by the electrical resistance of the joint faces. In this case, however, the electrodes are in the form of powerful vice clamps which hold the work-pieces and also convey the forging pressure to the joint [Fig.31].

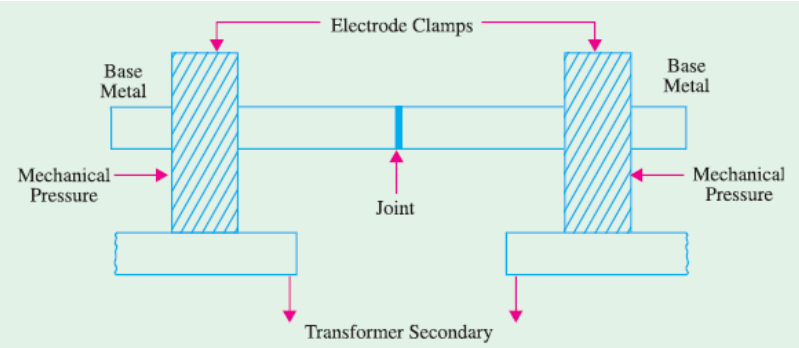


Fig-12

This process is useful where parts have to be joined end-to-end or edge-to-edge. i.e. for welding pipes, wires and rods. It is also employed for making continuous lengths of chain.

xxxxxxxxxxx

CHAPTER 4
ILLUMINATION

Light is a form of radiant energy. Various form of incandescent bodies are the sources of light and light emitted by such bodies depend upon the temperature of bodies. Heat energy is radiated into the medium by a body which is hotter than the medium surrounding it.

When the temperature increases the body changes red-hot to white-hot state, the wave-length of the energy radiated becomes smaller and enters into the range of the wave-length of light.

The ratio of the energy emitted by the body in the form of light to the total energy emitted by the body is known as the “radiant efficiency” of the body, which depends upon the temperature. Higher the temperature of the body; lower the wave-length of radiant energy and higher the efficiency.

Luminous Intensity:-Luminous intensity in any given direction is the luminous flux emitted by the source per unit solid angle, measured in the direction in which the intensity is required. It is denoted by symbol I and is measured in candela (cd) or lumens per steradian.

Lumen: - The lumen is the unit of luminous flux and is defined as the amount of luminous flux given out in a space represented by one unit of solid angle by a source having an intensity of one candle power in all directions.

i.e., Lumens= candle power × solid angle = $CP \times \omega$

Or, total lumens given out by source of one candela is 4π lumens

Illumination:- When the falls upon any surface, the phenomenon is called the illumination. It is defined as the number of number of lumens, falling on the surface, per unit area. It is denoted by symbol E and is measured in lumens per square meter or lux or meter-candela.

If a flux of F lumens falls on a surface of area A , then the illumination of that surface is

$$E = \frac{F}{A} \text{ lumens per meter}$$

Mean Horizontal Candle Power (MHCP):- It is defined as the mean of candle powers in all directions in horizontal plane containing the source of light.

Mean Spherical Candle Power (MSCP):- It is defined as the mean of candle powers in all directions and in all planes from the source of light.

Mean Hemi-Spherical Candle Power (MHSCP):- It is defined as the mean of candle powers in all directions above or below the horizontal plane passing through the source of light.

Brightness or luminance: it is defined as the luminous intensity per unit projected area of either a surface source of light or a reflecting surface and is defined by L.

$$L= \frac{1}{A\cos\theta} \text{candela/m}^2 \text{ or nits}$$

Solid Angle: Plane angle is subtended at a point in a plane by two converging straight lines and its magnitude is given by

$$\omega = \frac{Arc}{Radius} radians\ .$$

LAWS OF ILLUMINATION:

There are two laws of illumination (1) Law of inverse squares (2) Lambert's cosine law

1. LAW OF INVERSE SQUARES:

The law of inverse square states that "The illumination of a surface is inversely proportional to the square of the distance between the surface and the light source provided that the distance between the surface and the source is sufficiently large so that the source can be regarded as a point of source."

If a source of light which emits light equally in all directions be placed at the centre of a hollow sphere, the light will fall uniformly on the inner surface of the sphere, that is to say each square mm of the surface will receive the same amount of light. If the sphere be replaced by one of the larger radius, the same total amount of light is spread over a larger area proportional to the square of the radius. The amount which falls upon any square mm of such a surface will, therefore, diminish as the radius increases, and will be inversely proportional to the square of the distance.

Mathematically it can be proved as follows:

Let us consider surface area A1 and surface area A2 at distances r1 and r2 respectively from the point source S of luminous intensity I and normal to the rays, as shown in fig.

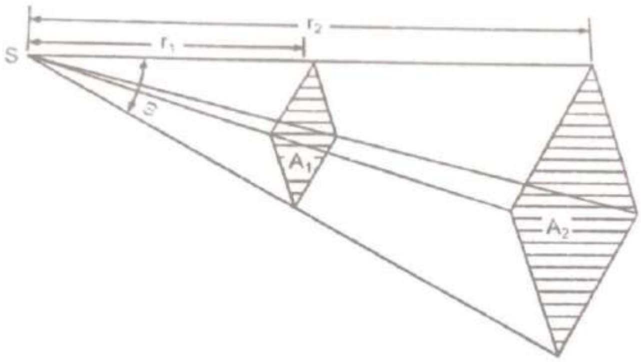


Fig. 1

Inverse Square Law:

Let the solid angle subtended be ω steradians

Luminous flux radiated per steradians = I

Total luminous flux radiated =Iω lumens

Illumination on the surface of area A1=Iω/A1 lumens per unit area

And area A1=ωr1

Illumination on the surface of area A1,

$E1=I\omega/\omega r1^2=I/r_1^2$ lumens per unit area

Similarly illumination on the surface of area A_2,

$E_2=I_w/A_2=I\ \omega/\omega r_2^2=I/r_2^2$ lumens per unit area.

2. Lambert's Cosine Law:

This law states that the illumination at any point on a surface is proportional to the cosine of the angle between the normal at that point and the direction of luminous flux.

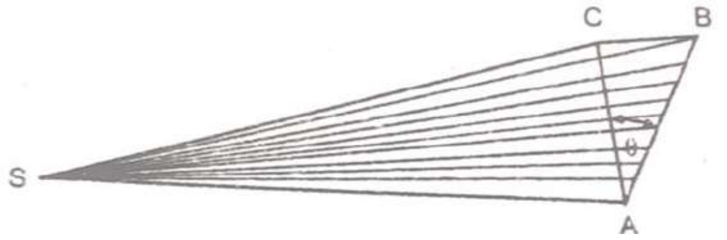


Fig. 2

Lambert's Cosine Law

The above figure shows that the area over which the is spread is then increased in the ratio

$$AB/AC=1/\cos\theta$$

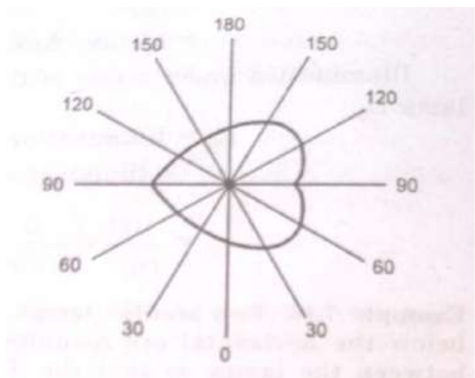
And the illumination decreases in the ratio $\cos\theta/1$

The expressions for the illumination then becomes

$$E=I\cos\theta / r^2.$$

POLAR CURVES:

The luminous intensity in all directions can be represented by polar curves. If the luminous intensity in a horizontal plane passing through the lamp is plotted against angular position then this curve is known as horizontal polar curve. If the luminous intensity in a vertical plane is plotted against the angular position, then curve is known as vertical polar curve. The vertical and horizontal polar curve is shown as fig.



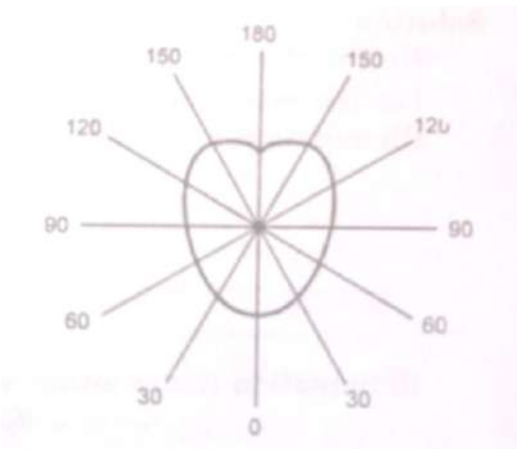


Fig.3a.Polar Curve for Horizontal plane

Fig 3.b. Polar Curve for Vertical Plane

The polar curves are used to determine the mean horizontal candle power and mean spherical candle power .These are used to determine the actual illumination of a surface by employing the candle power in that particular direction.

Maintenance Factor: The ratio of illumination under normal working conditions to the illumination when the things are perfectly clean is known as maintenance factor.

Illumination under normal working conditions / illumination when everything is perfectly clean.

Depreciation Factor: It is defined as the ratio of initial meter candles to the ultimate maintained meter candles on the working plane. It is also the inverse of the maintenance factor. Its value is more than 1.

TYPES OF LIGHTING SCHEMES:

The distribution of the light emitted by lamps is controlled by means of reflectors and translucent diffusing screens. The interior lighting schemes is classified as (a) direct lighting (b)semi-directing lighting (c) in directing lighting (d) general lighting.

Direct lighting: It is the most commonly used type of lighting scheme. In this scheme more than 90 percent of total light flux is made to fall directly on the working plane with the help of deep reflectors. It is mainly used for industrial and general outdoor lighting.

Semi-direct lighting: in this lighting scheme 60 to 90 percent of the total light flux is made to fall downwards directly with the help of semi direct reflectors, remaining light is used to illuminate the ceiling and walls. Such a lighting scheme is best suited to rooms with high ceilings where a high level of uniformly distributed illumination is desirable.

Semi-indirect lighting: In This lighting scheme 60 to 90 percent of total light flux is thrown upwards to the ceiling for diffuse reflection and the rest reaches the working plane directly except for some absorption by the bowl. This lighting scheme is with soft shadows and glare free. It is mainly used for indoor light decoration purposes.

Indirect Lighting: In this light scheme more than 90 percent of total light flux is thrown upwards to the ceiling for diffuse reflection by using inverted or bowl reflectors. in such a system the ceiling acts as the light source, and the glare is reduced to minimum. The resulting illumination is softer and more diffused, the shadows are more prominent and the appearance of room is more improved over which that results from direct lighting. it is used for decoration purposes in cinemas, theatres and hotel etc. and in workshops where large machines and other and obstructions would cause troublesome shadows if direct lighting is employed.

General Lighting: in this scheme lamps made of diffusing glass are used which give nearly equally illumination in all directions.

GAS DISCHARGE LAMP:

The basic principle of a gaseous discharge lamp as shown in fig. Gases are normally poor conductors at atmospheric and high pressures. When application of suitable voltage, known as ignition voltage across the two electrodes, as result in a discharge through the gas which is accompanied by electromagnetic radiation. The wave-length of this radiation depends upon the gas, its pressure and the metal vapour used in lamp.

Once the ionization has commenced in the gas, it has a tendency to increase continuously accompanied by a fall in the circuit resistance. In order to limit the current to a safe value of a choke or ballast is made. The choke performs the dual functions of providing the ignition voltage initially and limiting the current. Since due to use of choke the power factor becomes poor, i.e.0.3-0.4 .Therefore in order to improve the power factor of the gaseous discharge lamp use of a condenser. The colour of the light obtained depends upon the nature of the gas or vapour used.

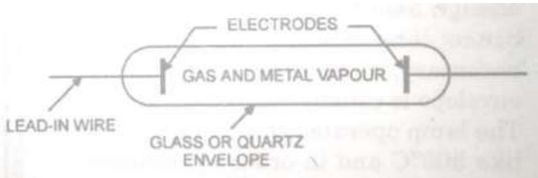


Fig. 4. Gaseous Discharge Lamp

The production of light by these lamps is based on the phenomenon of excitation and ionization in a gas or vapour . We shall now briefly discuss this phenomenon with reference to the structure of an atom.An atom has a positive nucleus and one or more electrons revolving around it in certain fixed orbits.in certain solids and gases here are what are known as free electrons which can escape from the influence of the nucleus of one atom and go over to another atom.There are thus a number of electrons which are mobile in nature.If a potential difference is applied to two electrodes placed in a gas having a large number of free electrons, these electrons will be attracted to the positive electrode and the velocity acquired by an electron will depend upon the potential gradient. During its motion towards the positive electrode, an electron will strike other atoms and one or more of the following results may be produced.

- ELASTIC COLLISION

The electron may be bounced off the atom it strikes and there may be no change in its velocity. This happens when the striking electron has a small amount of kinetic energy.

- EXCITATION

If the electron has acquired kinetic energy above a certain critical value in the process of passing through a certain potential which is termed as the excitation potential, the collision may cause one of the electrons to jump from its normal orbit into another one. This happens when the colliding electron has a kinetic energy of 2.1eV.The colliding electron imparts its kinetic energy to the atom that it strikes and this atom is said to be in an excited state.In this way the atoms can be placed in the 1^{st},2^{nd},3^{rd},4h or higher excited states depending upon the kinetic energy of the colliding electron.

- IONISATION BY COLLISION

If the kinetic energy of the colliding atom is large, it will completely knock out an electron from its orbit and this electron will now behave like a free electron and may produce more free electrons by collision. A large number of free electrons thus produced constitute a heavy current and an electric arc may result. This phenomenon is called *ionization*. Ionization potential is the potential difference through which an electron must travel to acquire energy for ionization by collision.

Neon Lamp

These belong to the cold-cathode category. The electrodes are in the form of iron shells and are coated on the inside. The colour of the light emitted is red and these lamps are mostly used for electrical advertising. High voltage is used for starting. If helium gas is used for in place of neon, pinkish white light is obtained. Helium and neon through coloured glass tubing produce a variety of effects. Figure below shows a circuit for a neon lamp. The transformer has a high leakage reactance, which stabilize the arc in the lamp. A capacitor is used for power factor improvement.

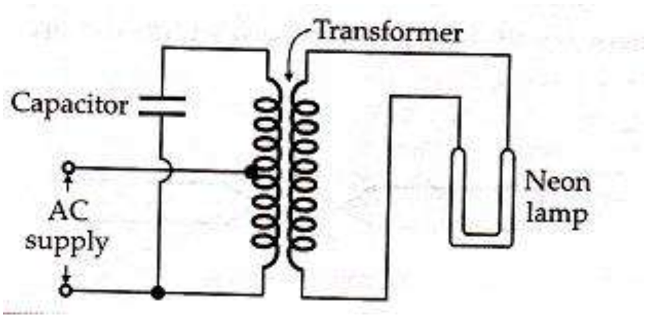


Fig. 5. Neon Lamp

Sodium Vapour lamp

Sodium vapour has the highest theoretical luminous efficiency and gives monochromatic orange-yellow light.The monochromatic light makes objects appear grey. Such lamps on account of this factor are used only for street and highway lighting.

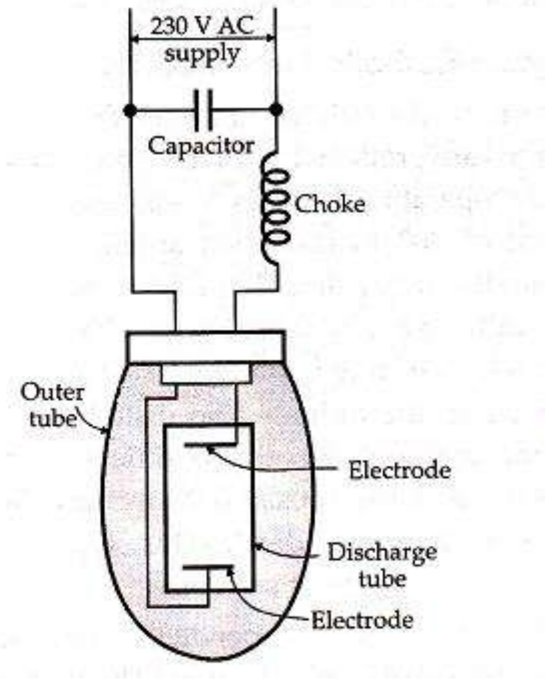


Fig. 6 Sodiun Vapour Lamp

The Lamp consists of a discharge tube having special composition of glass to withstand the high temperature of the electric discharge. The discharge tube is surrounded by an outer tube as shown below. For heating the cathode a transformer is included.Sodium below $60^{0}C$ is in solid state.For starting the lamp the electric discharge is allowed to take place in neon gas.The temperature inside the discharge tube rises and vapourises sodium.Operating temperature is around $230^{0}C$ .It takes about 10 minutes for the sodium vapour to displace the red colour of neon by its brown yellow colour. The lamp takes about half an hour to reach full output. A choke is providing for stabilizing the electric discharge and a capacitor for power factor improvement. The light output is about 40 to 50 lumens per watt.

Mercury Vapour Lamp

It is similar in construction to the sodium vapour lamp. The electrodes are tungsten coils containing an electron emitting material which may be a small piece of thorium or an oxide mixture. Argon is introduced to help start the lamp. The electric discharge first takes place through argon and this vaporizes the mercury inside the discharge tube. The electron emitting material supplies electrons to maintain the arc.

The space between the two bulbs is filled with an inert gas. The pressure inside the discharge tube may range from one to ten atmospheres in lamps used for lighting purposes as at these pressures the radiation is in the visible spectrum. If the pressure inside the discharge tube is low, most of the light is in the ultraviolet rgion.The efficiency is 30 to 40 lumens per watt.

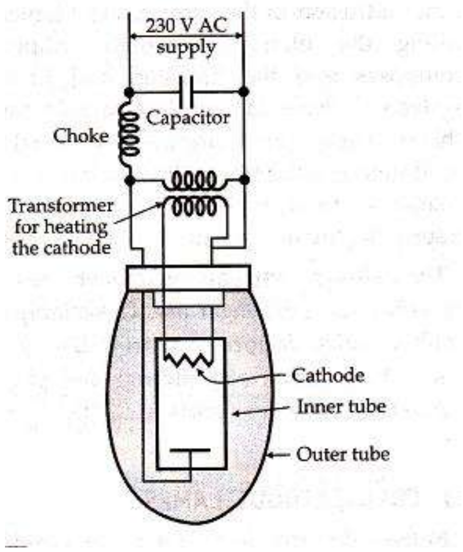


Fig.7 High pressure Mercury vapour Lamp

Fluorescent Lamp

In the mercury vapour lamp considerable amount of radiation is in the ultra violet range. By coating the inside of the tube by phosphor this ultra violet radiation is converted in visible light. Phosphors have definite characteristic colours but when mixed together they produce a large variety of colours. These phosphors are stable compounds and give a high output throughout the life of the lamp.

There are three types of fluorescent lamps:
1. Iron cathode or cold cathode type
2. Tungsten cathode, pre-heated type
3. Tungsten cathode, cold

In the cold cathode discharge tube under normal operating conditions which depend on the type and pressure of the gas and the type of electrodes,a glow discharge takes place which is discontinuous near the cathode where crockes and faraday dark space occur due to the formation of space charges in the gas.There is fairly large fall in voltage in this region.Then there is the positive column which provides useful illumination.The voltage drop along the positive column is proportional to its length.The large voltage drop at the cathode is independent of the tube length and depends only on the cathode material and the gas pressure. It may be between 100 and 200 volts. If, therefore , a cold cathode tube were to be operated from the mains,it would be very inefficient since most of the voltage will be utilized in overcoming the cathode voltage drop. It becomes necessary to use high voltage for the economic operation of this type of lamp.Also the lamp is not efficient unless its length is considerable.However, cold cathode tubes are of smaller diameter and can give any shape which makes them suitable for display and advertisement purpose.

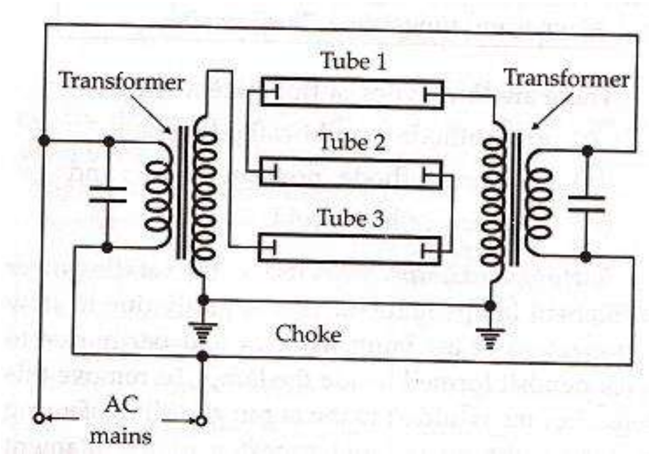


Fig. 8 Operation of the cold cathode lamps in series

Tungsten Cathode Preheated Type

In tungsten cathode preheated type electrons are produced by thermionic emission. Lower starting operating voltages are adequate. A transient voltage of 300 to 600 volts applied by the starter initiates the arc stream. The coating material decays in each starting of the lamp. The constant impact of electrons on the cathode also dislodges some of the emitting material. Finally so little of the materials is left that it is not possible to emit any electrons and the lamp becomes dead. This type of lamp is unsuitable for frequent starting

Fluorescent lamps produce flicker or stroboscopic effect since on 50 cycle supply they are extinguished 100 times a second. Single lamp cannot be operated without flicker. Flicker correction can be applied to pairs of lamps.

Radio interference is another effect produced by fluorescent lamps and has to be removed by suitable filter circuits. The advantage of fluorescent lamp is that its efficiency and life under normal conditions are almost three times those for filament lamps. The quantity of light obtained is superior, glare is minimum and the fluorescent light source casts soft shadows. However, the initial cost of the lamp and filling is higher than the incandescent lamps.

Starters of automatic starting switches are of two types:

1. Thermal Type
2. Glow discharge Type

The thermal starter has a heater coil which heats a bimetallic switch. The heater coil remains energized to keep the bimetallic switch open throughout the operation .It therefore, consumes a small amount of power. Figure below shows the circuit diagram of fluorescent lamp started by a thermal starter. When the supply is switched on the contacts of the bimetallic switch are closed and the current passes though the electrodes and heats them up. But after an interval of a few seconds the heater coil heats up the bimetallic strip and the bimetallic switch contacts open. This starts a high voltage transient across the electrode due to the presence of the choke or ballast in the circuit. An arc is struck between the electrodes due to the high voltage transient. The identical circuit showing the use of a glow starter can also be used as shown. The glow starter is enclosed in a glass bulb filled with neon or argon. One of the electrodes is a bimetallic strip.

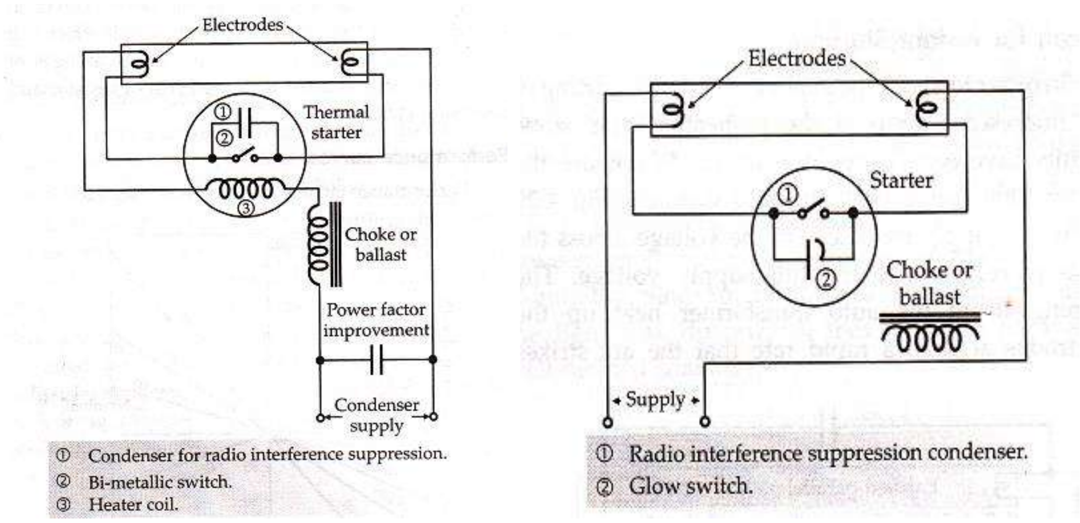


Fig.9 Fluorescent lamp

When the normal voltage is applied to the lamp, a glow discharge takes place across the glow switch and small current flows through the electrodes. The bimetallic strip expands due to the heating effect of the current in the glow discharge. The expansion of bimetallic strip causes the electrodes touch each other and the electrodes get pre heated due to flow of an appreciable amount of current. Mean while the bimetal cools. The glow switch opens and the resultant high voltage transient starts the arc discharge through the tube.

XXXXXXXXX

CHAPTER-5
INDUSTRIAL DRIVES

Introduction

Electrical energy is finding increasing application in industrial and commercial fields. Electric drive for industrial purposes is now almost universal. There are number of inherent advantages that the electric drive possesses over other forms of conventional drives. It is cleaner, more easily controllable and more flexible. With greater advancement in the development of Electric motors and control gear, the trend in the industry is towards an all-electric drive.

Both d.c. And a.c. is used for electric drives. Use of d.c. is limited on account of permissible voltage drop in feeders. But d.c. systems are still in use for many reasons.

The electric drive due to various inherent advantages has been universally adopted by the industry. Both A.C. and D.C. motors are used, however, A.C. system is preferred. The utilization of electrical energy is always advantageous as it is cheaper ,it can be easily transmitted at comparatively low line losses it is easy to maintain the voltage at consumer premises within the prescribed limits and it is possible to increase or decrease the voltage without appreciable loss of power.

In spite of the advantages of A.C. system, following are the applications of D.C. Industrial drives:

(i) For traction purposes, as in such application a very high staring torque is required. The starting torque can be obtained from D.C. series motor at low operating cost.
(ii) The speed of A.C. motors is almost constant, where as it can be varied easily in case of D.C. motor. Thus for variable speed applications such as lift and Ward Leonard system etc., the D.C. motors are preferred.
(iii) D.C. motors are also used in industry where very high accuracy of speed control is required.
(iv) The cost of change-over from d.c. to a.c. involves changes both in the power system and the consumer's equipment and is likely to be expensive.
(v) In some processes, Example: - electro-chemical, battery-charging etc. d.c. is the only type of power that is suitable.

Group Drive

Where a number of machines are driven through belts from a common shaft, it is known as *group drive*. Alternatively, each machine may have its own driving motor, in which case it is called individual *drive*.

In group drive case, one motor is used as a drive for two or more machines. The motor is connected to a long shaft. The machines are connected to this shaft through belt and pulleys. The use of this kind of drive is restricted due to the following reasons:

(i) If at certain instance all the machines are not in operation, then the motor will be working at low capacity.

(ii) In case of fault in the motor all the machines connected to this motor will cease to operate thereby paralyzing either complete or part of industry up till the time the fault is removed.

(iii) It is not possible to install any machine at a distant place.

(iv) The possibility of installation of additional machines in an existing industry is limited.

However, there are certain advantages of the group drive, which are detailed below:

i. Initial cost of installing the industry is low. For example , if the power requirement of each machine is 10 H.P. and there are 10 machines in the group, then the cost of ten numbers 10 H.P. motors will be much more than one 100 H.P. motor. Further, it is learnt from practical experience that the combined requirement of all these ten machines at a time will be less than 100 H.P. This further reduces the initial cost.

ii. In certain industrial processes one process is connected to another process and will be advantageous if all these interconnected processes are stopped simultaneously.

Individual drive :

In this case there is a separate driving motor for each machines. Such a drive is very common in most of the industries. It has the following advantages :

i) If there is a fault in one motor, the effect on the production or output of the industry will not be appreciable.
ii) Machines can be located at convenient places.
iii) Continuity in the production of the industry is ensured to a higher degree.

Following is the disadvantage:

i) Initial cost will be high.

Selection of Motors :

Due to the universal adoption of electric drive, it has become necessary for the manufacturer to manufacture motors of various designs according to the suitability and use in various classes of industry. This has resulted into numerous types of motors. For this reason, the selection of motor itself has become an important and tedious process. Taking into account the conditions under which a motor is required to operate, following factors will decide the type of motor required. :

1. Electrical Characteristics: The following are the electrical characteristics:

a. Starting characteristics
b. Running characteristics

c. Speed control
d. Braking

2. Mechanical Characteristics. These are:

a. Structural feature i.e. type of enclosure and bearing.
b. Method employed for transmission of power.
c. Noise.
d. Type of cooling.

3. **Size and Rating of motors.**

Following are the sub-heads under these characteristics;

a. Rating of the motor.
b. Suitability of the motor for continuous intermittent or variable loads.
c. Over load capacity.

4. Cost;

a. Initial cost.
b. Running cost.

In addition to the above factors, the type of current is also to be taken into consideration. From above it will be seen that the basic problem is to study carefully and thoroughly the load requirement, its surrounding and type of job it has to perform and then a motor which has the required characteristics and fulfil all the requirements is selected. The factors described above have been discussed in the following pages in detail.

Starting characteristics

The starting torque exerted by a motor should be large enough to accelerate the motor and its load to the rated speed in a reasonably short time. Some motors mat have to start against full load torque, Ex:- motors driving grinding mills or oil expellers. In the case of lifts and hoists, the motors have to start frequently with acceleration.

At the time of starting a motor two torques come into play: the torque required to overcome the static friction and the torque necessary to accelerate the motors and its load to the desired speed. The torques required for static friction cannot be easily determined. The torque for acceleration depends upon the load torque itself. The load torque may:

(i) Increase with speed i.e., may be proportional to $(\text{speed})^2$ as in the case of a fan or centrifugal pump OR
(ii) Remain constant with speed as in the case of a hoist.

The starting gear should, therefore be able to carry the starting current taken by a motor to a safe value consistent with the production of the necessary starting torque.

Starting Torque of DC Motors

Starting Torque in case of DC motors. Consider P poles motor producing flux $\emptyset$ webers per pole and let, I_a be the total armature current.if the number of parallel paths are A, then the gross torque T_g is given as:

$T_g = \frac{1}{2\pi} \times \frac{\emptyset ZPI_a}{A}$ m.Nw =0.159 $\frac{\emptyset ZPI_a}{A}$m.Nw =0.0162$\frac{\emptyset ZPI_a}{A}$ m.Kg

Now, whole of this torque developed will not be available at the pulley or is not available for doing useful work, since some of the power (Torque) developed is utilized in supplying friction and windage losses. The difference of gross Torque and the Torque lost in friction is called the shaft torque.

Let,ω be the angular speed of motor.

Power developed by the armature = $T_g\omega$ metre-Newtons or joules or watts

$$= \frac{2\pi NT_g}{60}$$

But 1 H.P.(metric) = 735.5 watt

$\therefore$ H.P.(metric) developed by the armature = $\frac{2\pi NT_g}{60\times 735.5}$

The torque therefore, depends upon the product of flux and armature current and is independent of speed i.e. , $T \propto I_a$.

In the case of a shunt motor, both the armature and the field are connected in parallel across constant voltage mains. The current taken by the field is, therefore, constant and hence the flux will be maintained constant so long as the field current remains constant. Therefore the torque in a shunt motor varies as the armature current. The torque –armature current curve is a straight line passing through the origin. Full-load current will produce full-load torque and twice the full-load current will produce twice the full-load torque.

In the case of D.C. series motors, the current in the series winding and the armature is same. The flux is dependent directly on the value of the current the motor draws. Torque is, therefore, proportional to the square of the armature current i.e. $\propto I_a^2$ .The torque-current curve is ,therefore, a parabola.But the flux varies as the current only upto the limit of saturation of the magnetic circuit and the torque current curve is parabolic in shape only up to the limit of saturation.Beyond the saturation point since”$\emptyset$ “does not vary appreciably the torque current curve is almost a straight line. A d.c. series motor is,therefore suitable for drives starting with heavy loads,Ex- electric train ,hoists and lifts etc.

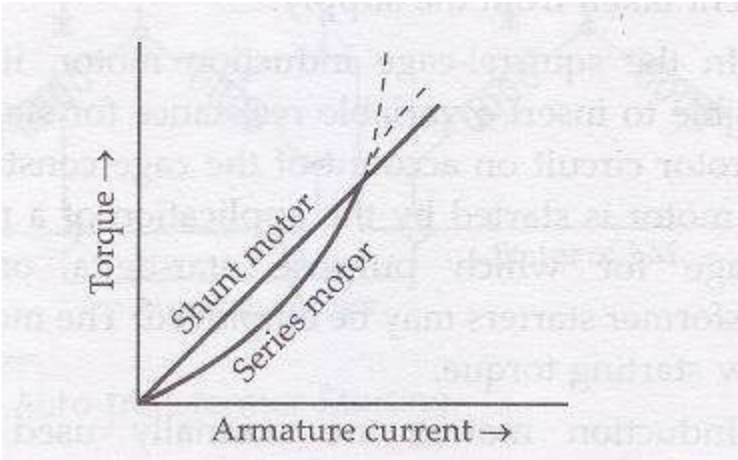


Fig..1 Torque Current characteristics of DC motor

Three-Phase Induction motors

In a three-phase induction motor, if

r_1 = stator resistance per phase
r_2' = rotor resistance per phase referred to the stator
x_1 = stator reactance per phase
x_2' = rotor reactance per phase referred to the stator
s = slip
V = stator applied voltage per phase,
Then the torque, T, is given by

$$T = k\,\frac{V^2 r_2'/s}{(r_1+\frac{r_2'}{s})^2+(x_1+x_2')^2} \quad \text{and}$$

Stator current per phase is: $I = \frac{V}{\sqrt{(r_1+\frac{r_2'}{s})^2+(x_1+x_2')^2}}$

If k is made unity ,the torque is expressed in synchronous watts per phase.
At starting,
s = 1
Therefore, starting Torque,

$$T_s = \frac{V^2 r_2'}{(r_1+r_2')^2+(x_1+x_2')^2} \text{ synchronous watts per phase}$$

and starting current per phase

$$I_s = \frac{V}{\sqrt{(r_1+r_2')^2+(x_1+x_2')^2}} \text{ amperes.}$$

The starting torque is a maximum if the rotor resistance per is made equal to its leakage reactance. It is, therefore, usual to start a slip ring induction motor with a variable resistance in its rotor circuit to have a good starting torque and to cut the resistance in steps as the motor speeds up. The resistance in the rotor circuit also serves the purpose of limiting the starting current taken from the supply.

In the squirrel cage induction motor, it is not possible to insert a variable resistance for starting in the rotor circuit on account of the cage construction. The motor is started by the application of a reduced voltage for which purpose star-delta or auto-transformer starters may be employed. The motor has a low starting torque

Induction motors are normally used where constant torque is required, ex- in paper machinery, textile machinery, compressors, conveyors etc. Squirrel cage motors are more reliable, cheaper and easier to use where as phase wound motors are expensive and maintenance is complicated. The former are used for low and medium H.P. while the latter are used for high H.P.

Motors with double cage have a high starting torque. The outer cage is made of high-resistance metal bars and inner cage is made of copper bars. The inductance of the inner winding is higher than that of the outer high resistance winding. At the instant of starting, the motor induced currents are at the line frequency and the inner cage has a high reactance ($2\pi fL$) with the result that the rotor currents remain confined to the outer cage despite its high resistance. The starting torque is, therefore, high.During normal running, the reactance of the inner cage decreases ($2\pi sfL$), and the rotor currents are now confined to the inner cage which is a low resistance winding. This gives a high efficiency of the motor.

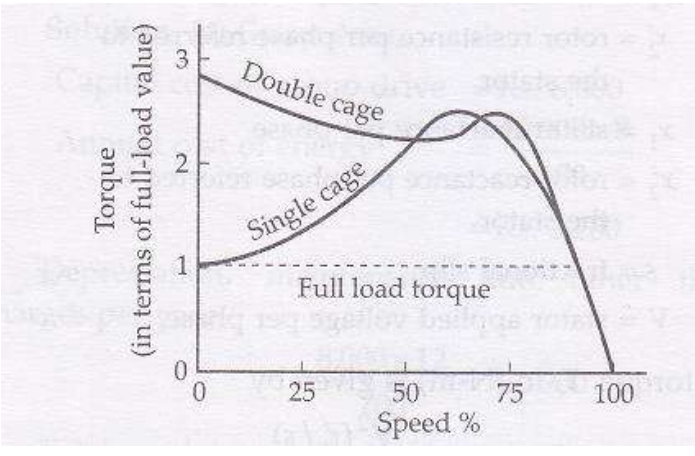


Fig.2 Torque speed characteristics of Induction motor

The Figure above shows the speed-torque curves of a single cage and double cage motor.

An important relation existing for three-phase induction motor, i. e.

$$\frac{\text{Starting Torque}}{\text{Full} - \text{Load torque}} = \left(\frac{Starting\ Current}{Full - Load\ Current}\right)^2 \times Full\ Load\ slip$$

$$i.e., \frac{T_s}{T_L} = \left(\frac{I_s}{I_L}\right)^2 \times s_{FL}$$

Methods of starting 3-phase induction motors

It is desirable to start a.c. motors at full voltage to attain simplicity and economy in the starting gear. Induction motors can be designed and built to enable them to be started on full voltage. But in case of cage motors the starting current may be large enough to produce considerable voltage drop in the distribution system which may adversely affect other apparatus and also cause light flicker.

Induction Motors are therefore, started on reduced voltage .The various methods of starting are discussed as follows:

1. **Resistor starting method**

A series resistor is used in each line and may be arranged in a manner that the resistance is reduced to zero in steps so that the motor current may increase to the full value gradually and transients are avoided.

The torque efficiency which is

$$\frac{\text{Torque developed by the Motor} / \text{ Full load torque}}{\text{Current of Motor } / \text{ Full load current}}$$

A motor having a starting torque equal to twice the full-load torque and a starting current six times the full load current ha s a torque efficiency = 2/6=0.33.

If the same motor is started at 60 percent voltage by using resistors, the line current will be(0.6 x 6 x full-load current) and the starting torque will be [0.6 x 0.6 x 2 x full-load torque].The Torque efficiency will be (0.6 x 0.6 x2)/0.6 x 6 = 0.2 .

2. **Reactor starting**

It is not a frequently employed as resistors or auto transformer starting, though the method is similar to resistor starting. The acceleration is very smooth in this case though the acceleration time is comparatively longer.

3. **Autotransformer starting**

Taps are provided on the auto transformer so that the motor can be started at reduced voltage. Taps are for 50, 65 and 80 percent of line voltage; the 50 percent tap being provided only in the case of sizes above 50 H.P. In the second method of operation the large transient current is reduced since the motor is always connected through the auto-transformer winding to the line.

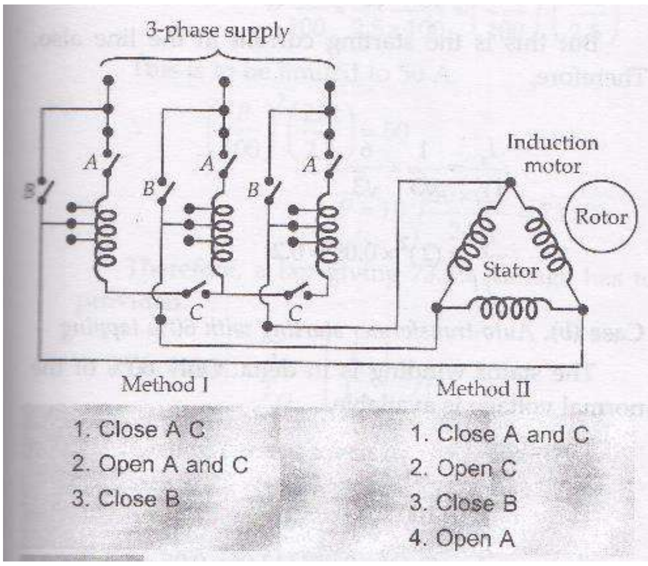


Fig. 3. Starting of SQIM by Autotransformer starter

We can determine the reduction in the starting current when using auto-transformer starter. Consider an auto-transformer with a transformation ratio $K = \frac{\text{Primary Voltage}}{\text{Secondary voltage}}$. Consider also that the motor has a starting torque equal to twice full-load torque and starting current equal to six times the full load current. If the motor is started at full voltage, phase voltage $E_{ph} = E/\sqrt{3}$ and starting current = 6I.when the motor is started through an auto-transformer, the phase voltage is E_{ph}/k and the starting current = 6I/k.

Also we have $\frac{E_{ph}}{V_{ph}} = k = \frac{N_1}{N_2} = \frac{I_2}{I_1}$

Where N_1 =No. of primary turns
N_2 =No. of secondary turns
I_1 = primary current
I_2 = secondary current
$I_2 = 6I/k = kI_1$ or $I_1 = (6/k)^2 I$.

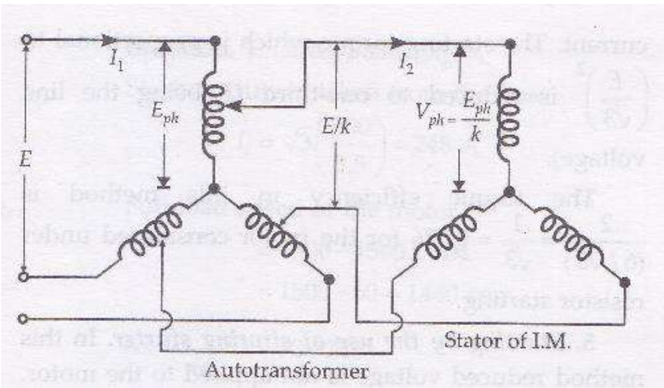


Fig..4 Autotransformer starting principle

Therefore, the line current is reduced inversely as the square of the ratio of transformation. Since the torque is proportional to the square of the applied voltage, the starting torque is proportional to V_{ph}^2 or $(E_{ph}/k)^2$. If T is the full voltage starting torque, starting torque at reduced voltage = T/k^2.

$$\text{The torque efficiency} = 2/(6/k)^2 = 0.33\ k^2$$

4. Star-delta starting

The stator of the cage motor is connected in star in the starting position and in delta in the running position, so that $\frac{1}{\sqrt{3}}$ of the line voltage is impressed on each phase at the time of starting.A star-delta starter is shown schematically in figure …, The starting line current of the motor with star-delta starter is also reduced to $\frac{1}{\sqrt{3}}$ full voltage starting line current. The starting torque, which is proportional to $(\frac{E}{\sqrt{3}})^2$ is reduced to one-third. (Where E is the line voltage).

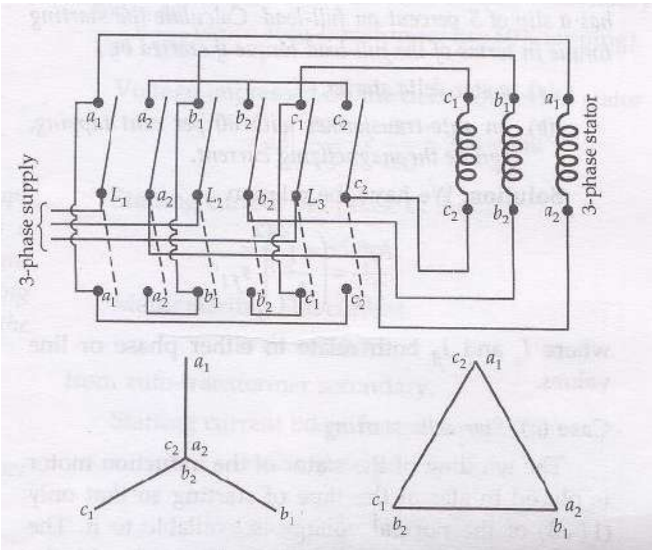


Fig..5. Star delta starting

The torque efficiency in this method is $\frac{2}{6/\sqrt{3}} = \frac{1}{\sqrt{3}} = 0.576$ for the motor considered under resistor starting.

5. **Starting by the use of slip-ring starter** :

In this method reduced voltage is not applied to the motor. The full voltage is applied to the stator but resistance is inserted in each phase of the wound rotor. Since the stator and rotor can be regarded as the primary and secondary of a transformer, the resistors in the secondary limit the currents in the rotor winding and since in a transformer $\frac{I_1}{I_2} = \frac{V_2}{V_1}$,the stator current I_1 is also reduced.

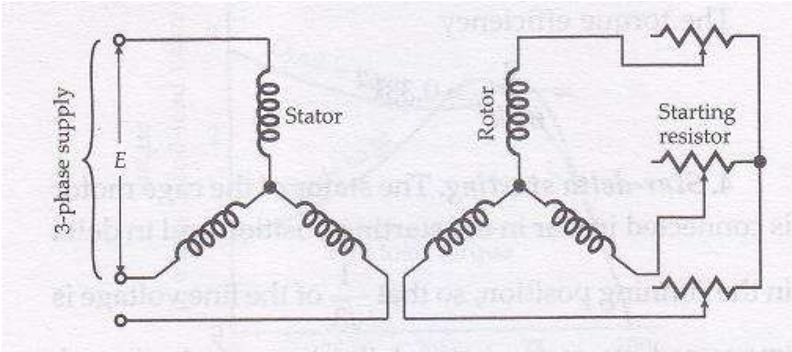


Fig.6. Rotor resistance starter

Limitation of size

From the foregoing examples it is seen that the starting torque in case of cage motor varies with the type of starting method employed. In fact the starting difficulty of cage-type motors has limited their application to loads requiring not more than about 40 or 50 hp with not more than about 50 percent full-load torque. The slip-ring motor is used beyond this limit.

Next, discussion will made in brief about the starting of other type of motors.

1.4. **Single-Phase Induction Motors**

A single-phase induction motor does not have a rotating magnetic field. it has only a pulsating field and therefore does not possess any starting torque. The following three methods are employed to make the motor self-starting.

(a) **Pole-shading**: The motor is of squirrel-cage type and the stator pole is shaded by a heavy copper wire or strip. The current induced in the shading coil causes the magnetic field through the shaded portion of the pole face to lag behind the main flux thereby producing a rotating Magnetic field. Such motors have a low starting torque but are quite economical in small sizes.

(b) **Phase-splitting**: A two-phase supply is obtained from a single-phase line by using a capacitor. The motor has a cage-rotor and a stator containing two separate windings located in the same manner as for a two-phase stator. One of the windings is connected directly to supply and the other through the capacitor. A rotating magnetic field is obtained in the air-gap. There are two types of capacitor motors: the capacitor start motor, in which case the capacitor is in circuit only during the starting period and is disconnected at a predetermined speed by a centrifugal switch; the other type is the capacitor start and run motor where the capacitor is connected permanently and improves the power factor of the motor.

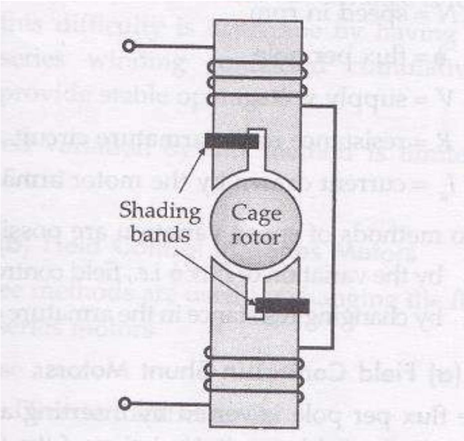


Fig..7 Shaded pole motor

(c) Repulsion motor starting: the rotor has a repulsion motor winding and therefore starts as a repulsion motor giving high starting torque. As it runs to speed a centrifugal device short-circuits the commutator bars and lifts the brushes, converting the motor into a plane squirrel cage one.

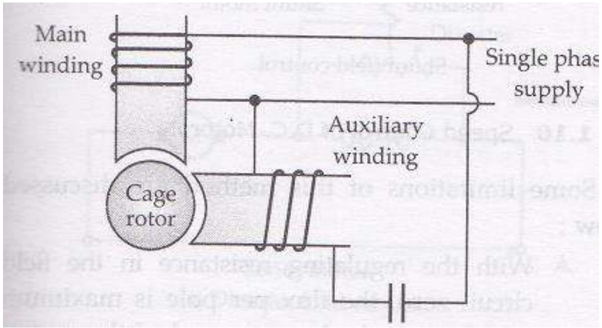


Fig..8 Split phase starting

(d) Synchronous motors: There is no starting torque in a synchronous motor. It has to be run up to synchronous speed by another motor and synchronized to the supply. To make it self-starting a cage winding is provided on the poles. It starts as a plain squirrel cage motor and when it has attained nearly synchronous speed at no-load, the d.c excitation is switched on and the rotor pulls into synchronism. Starting torque between 50 to 100 percent full load torque can be obtained with twice full load current. The synchronous Induction motor has a cylindrical rotor with a slip-ring induction motor winding. It starts as a slip-ring induction motor with good starting torque and when it has almost reached the synchronous speed, d.c. is passed through the rotor winding making rotor pull into synchronism.

A.C.Commutator Motors: These are started by the application of reduced voltage or by shifting the brushes.

Running Characteristics

The running characteristics of a motor include the speed-torque or the speed-current characteristics, losses, efficiency and power factor at various loads. Power factor consideration crops up in the case of a.c. motors only.

D.C. Motor

In the case of DC shunt motors speed is fairly constant with load; there is only a slight fall in speed as the load comes up. The speed torque characteristic is a slightly drooping straight line.

For the DC series motor the speed is normally high at low loads and decreases as the motor is loaded .The speed –Torque characteristics is a supply drooping curve.

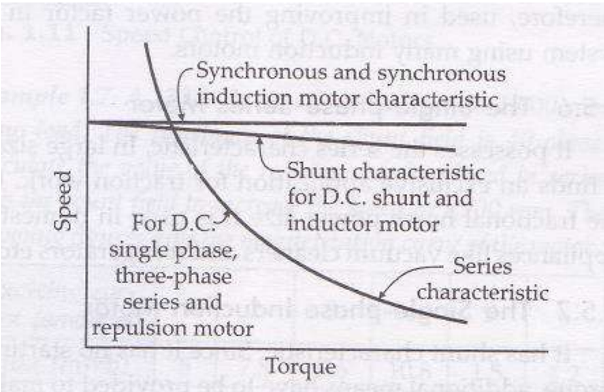


Fig..9 Torque speed relation of dc and ac motors

In the compound motor, the speed-torque characteristics may be made to lie anywhere between the pure shunt and the pure series by suitably adjusting the series and the windings.

The Theree-phase Induction Motors

It possesses shunt characteristics. The power factor is very poor at low loads but improves as the load increases. The power factor, however, always remains less than unity.

The Synchronous and synchronous-Induction motor

The synchronous motor is a constant speed motor: The speed is fixed by the frequency of the supply. It is not, however, self starting. It is started by an auxiliary motor and synchronized to the supply .This disadvantage is eliminated in a synchronous-Induction Motor where the machine starts as plane Induction motor and when the speed is very near the synchronous speed the DC excitation to the rotor is switched on and the motor pull into synchronism.

By varying the field excitation of these types of motors the power factor may be made unity or even be made leading. An over-excited synchronous motor works as a leading power factor while an under-excited motor works as lagging power factor. It is, therefore used in improving the power factor in a system using many induction motors.

Single-phase series motor

It possesses the series characteristics. In large sizes it finds an exclusive application of Traction work .In fractional horse power size it is used in domestic appliances like vacuum cleaners and refrigerators etc.

Single-phase Induction motor

It has shunt characteristics. Since it has no starting torque, additional means have to be provided to make it starting. Repulsion start and the capacitor start motor are the common modifications of the single phase induction motor.

Repulsion motor

It has series characteri9stics and closely resembles the series motor in construction.The armature is short circuited in itself.

SPEED CONTROL :

Control of speed for an industrial drive depends upon the nature of work being carried out. A certain operation may require a continuously varying speed; another one may only require two fixed speeds.Some times creeping speed may be necessary to adjust the work.For most industrial drives ,however,a control speed within $\pm$ 20 per cent may be suitable

Speed control of D.C. motors :

The speed of D.C. motors is given by the expression

$$N \propto \frac{V - I_a R}{\emptyset}$$

Or, $N = K \frac{V - I_a R}{\emptyset}$

Where, N = speed in rpm

ϕ= flux/pole

V = supply voltage
R = resistance in the armature circuit

I_a =current drawn by the motor armature

Two methods of speed variations are possible:

1. Flux variation or field control

2. By changing resistance in the armature circuit

Field control in shunt motors

The flux per pole is varied by inserting an extra resistance in the field circuit. Variation of the flux per pole changes the speed of the motor.

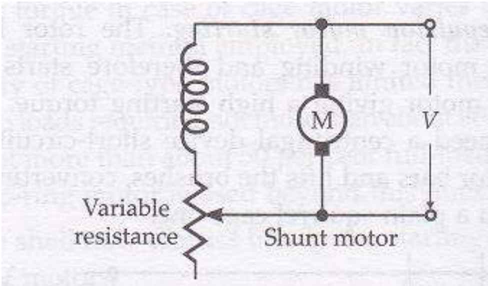


Fig.10 Shunt Field Control

Some limitations of this method are discussed below:

i. With the regulating resistance in the field circuit zero, the flux per pole, is maximum which gives the lowest speed of the motor. Motors are usually designed to work at a speed slightly less than the rated speed when the regulating resistance is zero. It is obvious that any lower speed than this cannot be achieved by this method.

ii. The speed of the motor N is proportional to $VI/\emptyset$ where as the full-load torque T is proportional to TN or VI_a which is constant. Therefore, this method can be utilized only where the horse power of the load remains constant.

iii. There is a limit to which the field can be weakened to obtain high speed. At such a speed the motor will tend to draw large current to develop the same torque. But this will result in the main field ampere-turns becoming much smaller than the armature mmf. The armature reaction will demagnetize and distort the main flux making the operation of the motor unstable. In motors where a wide speed-range is required, this difficulty is overcome by having a light series winding connected cumulatively to provide stable operation.

Sped variation by this method is limited to a ratio of 5:1.

Field control in series motors

Three methods are used for changing the flux per pole in series motors

These are :

- Diverter field control
- Tapped field control

- Series-parallel field control

i. **Diverter Field control** :

A shunt is employed in parallel with the series field to divert a part of the current in the series field thus causing field weakening. Speeds higher than normal are attained when the diverter is used.

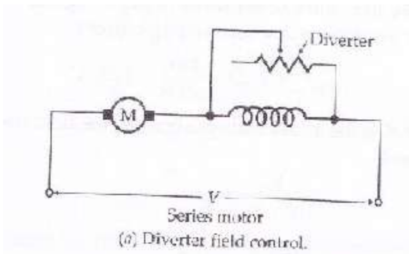


Fig.11 Field diverter control

Tapped Field control:

Tapping are provided on the field winding and current may be passed through different number of turns there by changing the field ampere-turns. This method is commonly used for series motors used in traction work.

Series-Parallel Field control:

The field winding is designed in two sections which may either be connected in series or in parallel .The field ampere-turns are reduced to half the value in parallel connection as compared to those in series connection. The speed, therefore, becomes about twice the initial value. Though the method is simple and inexpensive, only two speeds are possible.

<u>CONTROL OF SPEED BY CHANGE OF SERIES RESISTANCE IN THE ARMATURE CIRCUIT</u>

Since $N \propto \frac{V-I_aR}{\emptyset}$, the speed of a D.C.motor can be changed by varying R, resistance in the armature circuit. The torque of a motor is proportional to the product of the flux,$\emptyset$ and the armature current I_a .In the case of a shunt motor,since $\emptyset$ is constant,N will be proportional to $V-I_aR$ .if constant torque is required I_a should remain unchanged .But since speed is to be varied R has to be varied.Increase of R(for constant torque and therefore constant armature current)will give decreasing values of speed.The minimum value of R is R_a ,the resistance of the armature itself.The figure shows the armature speed torque characteristics.

$$R = r + R_a$$

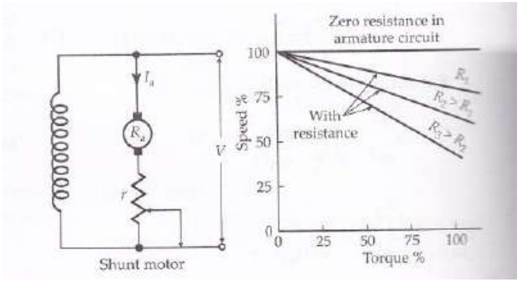


Fig.12 Speed control of shunt motor taking series resistance in armature

For a theoretical value of R = 0, the horizontal straight line passing through 100% speed ordinate is the limiting value. For any other values of R which may be $R_1 = r_1 + R_a$ or $R_2 = r_2 + R_a$, $R_2 > R_1$ or $R_3 = r_3 + R_a$, $R_3 > R_2$ etc. the curves are as shown.

If this method is used for a load requiring constant torque at all speeds, the armature current must remain constant and so the input to the motor (i.e., armature) is also constant. But the output decreases with the decrease in speed and hence the efficiency of the motor is poor at lower speeds. The power loss takes place in the controlling resistance r. In the case of fans and centrifugal pumps where the load torque decreases at lower speeds, this method may be quite convenient and economical for short periods. Creeping speeds may also be obtained by this method.

In a series motor, an increase in the armature circuit resistance will decrease both speed and torque. Since the flux is dependent on the armature current the torque is proportional to I_2. For a constant torque if different speeds are required, current (*I) has* to be constant which will make Ø constant. For reducing the speed resistance is to be increased.

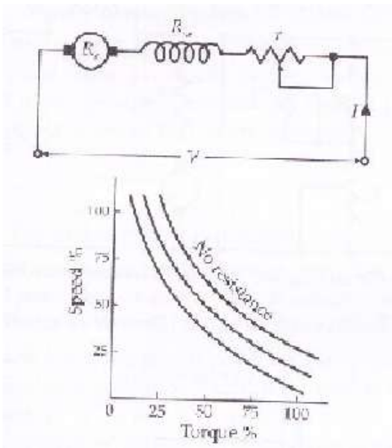


Fig.13 Speed control of series motor with series resistance in armature

Control of Motor Speed by Shunting the Armature by a Resistance

The arrangement of varying the speed of a d.c. motor by changing the series resistance in the armature circuit is at times not applicable as the speed of the motor rises if the load is reduced. We can see from the equation $N \propto (V - I_aR)/\emptyset$ that as I_a diminishes N increases. To eliminate this drawback, the armature is shunted by a variable resistance. A series resistance is also used as shown in fig. 1.14. By adjusting P and Q a number of speed torque curves can he obtained.

If we apply Thevenin's Theorem to the circuit in fig. 1.14(a), we get P and Q in parallel i.e., short circuit resistance R_{sh} by short-circuiting the source of supply and removing the branch (i.e., armature) through which we wish to find the value of the current flowing. Therefore the open circuit voltage across the armature is V_{oc}= {P/ (P+Q)} V. Fig. 1.14(b) shows the equivalent circuit based on The-venin's Theorem. The current is given by

$I=V_{oc}/ R_a+ R_{sh}$ Where $V_{oc}=P/ (P+Q)$ V and $R_{sh}=PQ/P+Q$ = short circuit resistance.

The efficiency of this method is poor and heavy currents may be drawn from the supply at certain speeds.

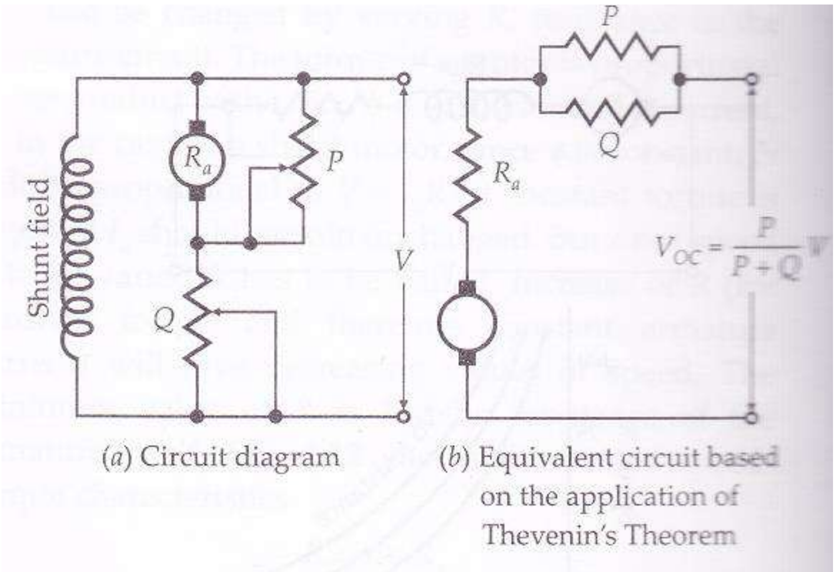


Fig. 14. Controlling resistance in parallel arrangement

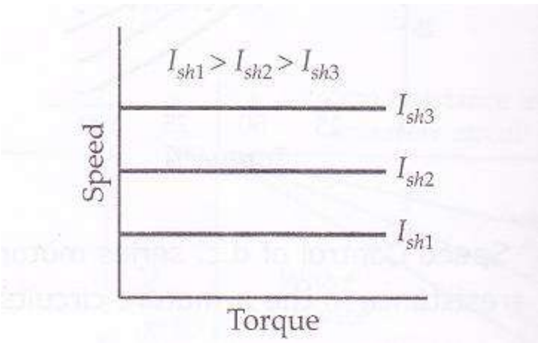


Fig. 15 Speed Torque characteristics

Booster Control

In Figure, M_1 is the main motor used for driving the load. It is excited from constant d.c. supply. B is the separately excited booster whose voltage and polarity can be controlled by the booster field BF. With the arrangement of tapping the voltage from the parallel resistances the current through the booster field can be reversed which reverses the polarity of the booster armature voltage. Thus, necessary boost or buck can be provided to M_1.M_2 is the shunt motor which drives the booster armature. The speed can be varied over a wide range depending on the size of the booster. This method is, however, suitable for small motors; otherwise the size of the booster becomes uneconomical.

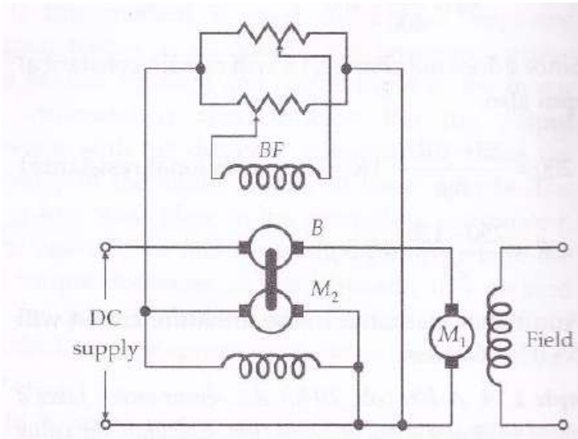


Fig.16 Booster control

Ward-Leonard System of Speed Control

In this method, the current in the motor armature is supplied from a variable voltage source, say a special motor generator set and the field is given a constant excitation so that in the expression

$$N=K\ V-I_aR/\emptyset$$

Where V is variable and $\emptyset$is constant.

This system is used where a very fine speed control is needed as in the case of colliery winders. Fig. 17 shows the application of this method.

There is a motor generator set whose motor M is a 3 phase slip ring induction machine coupled to the d.c. generator G. A flywheel is mounted on the motor generator shaft to equalize the fluctuations in the load. A d.c. exciter is also coupled to the M.G. set. This exciter supplies the excitations to the generator G and the load motor. In order to equalize the fluctuations in the load the flywheel must decelerate during the lean load period. This is achieved with the help of the torque motor which puts extra resistance in the slip-ring rotor by operating the liquid rheostat when peak load comes on. The torque-motor is supplied through the CT_S as shown and is actuated by current proportional to that drawn by the motor of the M.G. set. When slip-ring motor tends to draw heavier current, the torque-motor is actuated through the CT_S increasing the resistance of the liquid rheostat which decreases the speed of the M.G. set and allows the flywheel to give up a part of its stored energy. The whole process is reversed when the load decreases. The direction of the load motor can be reversed by reversing the polarity of the voltage supplied to it by the generator G. The polarity of generator voltage can be changed through the reversing switch which reverses the excitation of the generator. The Ward Leonard Control with a flywheel is known as Ward-Leonard-IIgner Control.

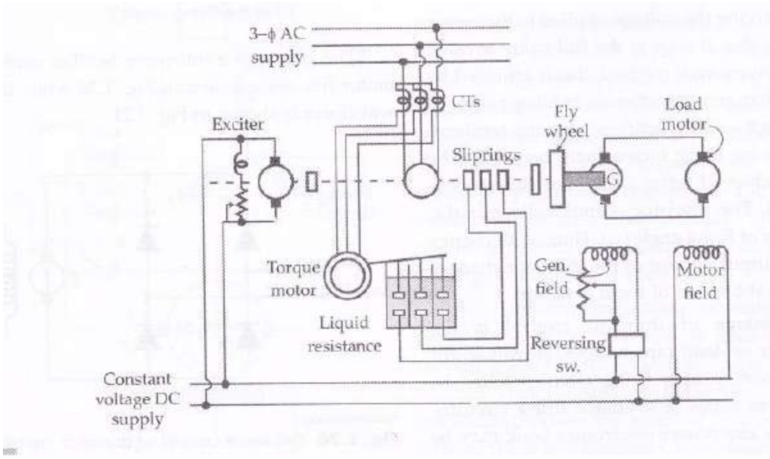


Fig. 17 Ward Leonard Ilgner control

SPEED CONTROL OF INDUCTION MOTORS

Speed control of Induction motor is given by the following equation:

$$N_r = N_s(1\text{-}s) = \frac{120f}{P}(1\text{-}s)$$

$[N_s$ =synchronous speed = 120f/P]

Thus, the speed of an induction motor, broadly speaking, can be regulated by three methods i.e. by varying any of the above quantities Viz. frequency, number of poles, slip.

Thus speed change can be achieved by varying the frequency. Induction motor has drawbacks of developing low starting torque, drawing heavy starting current and having no easy means of continuous easy speed control. On the other hand, assets of induction motor are: trouble free operation, less maintenance, high voltage operation consequently needing reduced amount of current and automatic regeneration. In addition to these, because of extreme mechanical simplicity in the construction, the rotor can resist centrifugal forces better than D.C. armature and therefore, for a given amount of iron and copper, more power can be produced.

FREQUENCY METHOD OF SPEED CONTROL

The above equation suggests that the speed of a induction motor is directly proportional to the frequency of supply voltage.

<u>Control from variable frequency supply</u>

Induction motor operates at a high efficiency and power factor at speeds near to its synchronous speed. The difference between actual speed and synchronous speed, which is called slip, represents losses in the rotor. Thus in induction motor, operating from constant frequency supply, slip has to be small if efficiency is to be high. In other words, motor should operate at high speed (near synchronous speed).If however, synchronous speed itself is brought down near to actual slow operating speed, motor will still be working at high efficiency. This is achieved in variable frequency supply. Another advantage of feeding low frequency supply to induction motor at starting is that it does not take heavy starting current. This is proved as follows:

If suitable variable frequency supply is made available, induction motor can develop high starting torque without excessive rotor currents when it is supplied with low frequency voltage supply say ½ to 9 cycles. Intersection of the stable region of torque speed curve with the load torque curve gives the operating speed. As the frequency of supply is reduced, torque curve shown dotted will move more towards left. This increases the starting torque(T_s).Another advantage of variable frequency supply is that as the motor speed falls ,the frequency of supply is reduced.

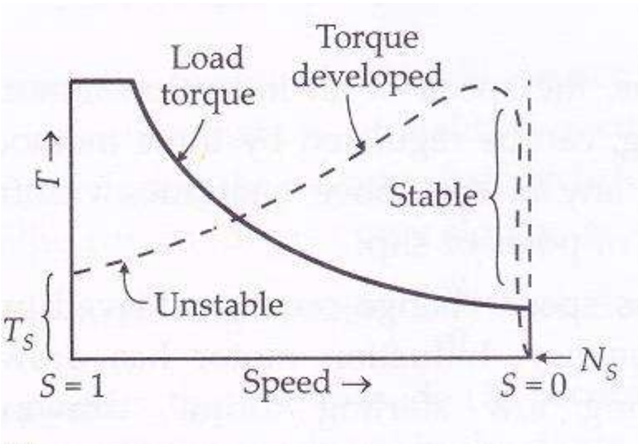


Fig.18 Electromagnetic torque and load torque profile

This avoids the operation of motor on unstable portion of the speed torque curve. Induction motors apart from the advantage of avoiding commutator maintenance, can be of smaller size as they can be robust higher speeds and higher temperatures than D.C. motors and in addition they are more efficient. Induction motor is excellent for the industrial purpose as it has excellent power weight ratio, great mechanical strength, suitability for higher speeds, no sliding contacts and high starting torque without overload. But all these advantages are available only through close control of flux and frequency.

Control by Variable Frequency inverter Employing Thyristor.

With a conventional rotary converter, it was not possible to obtain low frequencies say ½ to 9 cycles, so that attempt for developing high starting torque always resulted in heavy rotor currents. However with the development of silicon controlled rectifier (SCR), used as inverter, the frequency of three phase supply can be adjusted from 0 – 150 cycles.

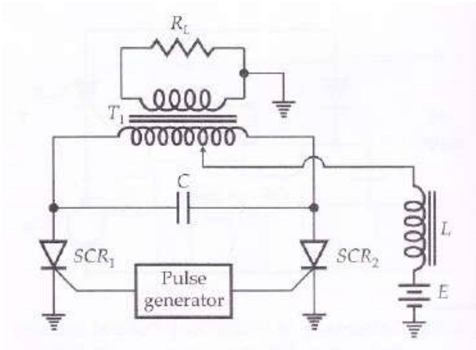


Fig. 19 Schematic of variable frequency inverter

Fig. 19 shows a single phase parallel inverter circuit using two SCR_S. to start with, assume SCR_1 as conducting and SCR_2 as off. Left plate of condenser C will be at negative potential and right plate will be at positive potential at double the voltage of d.c. source due to auto transformer action. A trigger from the pulse generator to SCR_2 will switch it on. Now condenser C will send current through SCR_2 and block SCR_1. D.C. source will send current through inductance L, transformer T and SCR_2. Current pulse flowing through transformer will produce polarity in the secondary of transformer of opposite sign to that produced previously when SCR_1 was conducting. Condenser C will now be charged with right band plate negative and left hand plate positive to double the d.c. source voltage. Now when SCR_1 is triggered, condenser C will again discharge but this time through SCR_1 to block SCR_2. In this way cycle repeats. For 1:1 turn ratio of transformer, the peak value of a.c. square wave is half the voltage of d.c. source and frequency of a.c. supply depends upon the frequency of the firing of SCR_S.

Pole Changing Method of Speed Control

This method is applicable to squirrel cage motors only, as their rotors can adjust themselves to any number of poles. This method of speed control is used for driving drilling machines which require different speeds for drilling into different metals. This is also used for lifts where regenerative braking can be applied by pole changing. On increasing the number of poles synchronous speed becomes less than the actual running speed and motor now works as an induction generator. Some motors have two stator windings, wound for different poles.

By Applying Variable Voltage to Stator

In Fig. 20, two speed-torque curves of a motor with different applied voltages are shown. Torque developed is proportional to square of applied voltage. Hence speed-torque curve ‘A’ will be for higher voltage and ‘B’ for lower voltage. Intersection of load torque line with the torque developed, gives us speed (N_1) with higher voltage applied and N_2 speed with lower voltage applied. With constant torque loads, speed control by this method gives limited variation of speed. However for loads whose torque varies as the square of speed, this method gives wide range of speed.

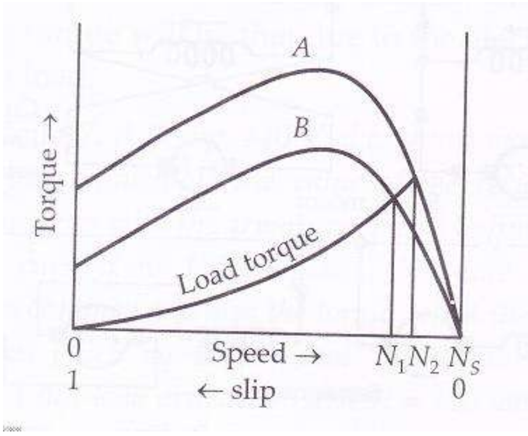


Fig.20 Torque speed characteristics with voltage

By Varying the Rotor Copper Losses

For constant load torque applications, slip is directly proportional to the rotor copper losses in induction motor. By increasing the rotor copper losses, slip will be increased and hence speed can be decreased. Rotor losses can be regulated by following methods.

1. **By Inserting External Resistance in the Rotor Circuit**.
In induction motor, for a given torque to be developed, rotor current remains constant. Therefore, if resistance is inserted in the rotor circuit, it will increase the rotor copper losses and, therefore, slip. This method, besides being wasteful for low speeds, requires heavy three phase controller to dissipate the losses.
2. **By Cascade Control**. Instead of wasting the energy in rotor resistance and creating a problem of disposal of heat, it can be taken out of the rotor and fed to another motor which is capable of taking power at low frequency. This motor is usually mechanically coupled to the main motor. Thus power taken from the rotor is converted to mechanical energy. This arrangement would give increased torque output at reduced speed. This rotor power at low frequency can also be converted to power at supply frequency by means of frequency converter and fed back to line. Arrangement of connecting auxiliary machine to the rotor of main motor to perform any of the above functions is called cascading.

Speed Control by Slip Coupling

This method allows driver shaft to run faster than the driven shaft by an amount which can be controlled. There will be some power loss in the coupling itself. One form of slip coupling consists of stator and rotor. Rotor is squirrel cage type and is keyed to the driven. Around the rotor is stator containing d.c. excited salient poles. Stator is keyed to the driver. In ordinary induction motor rotating magnetic field is produced electrically by the three phase currents. As against this, in slip coupling, magnetic field is mechanically driven by the driver. This will exert torque on the rotor which will be transmitted to the driven. In induction motor the magnitude of the torque is dependent upon the magnitude of the rotating field or applied voltage. In slip coupling, also the magnitude of the torque at which slip occurs can be varied by the excitation of the slip coupling which is conveyed to it through two slip rings. Torque slip characteristic of slip coupling is essentially same as that of induction motor. Slip coupling can, therefore, make it possible to have variable speed drive from constant speed driver.

ELECTRIC BREAKING

In many industrial drives, it becomes important to stop motor and its work in a reasonably short time, as in the case of planer where the tool must be stopped quickly at the end of its stroke. To achieve this breaking system has to be used.

Two types of breaking systems are possible:

1. Mechanical or friction braking where the motor is stopped by using a brake shoe or band on brake drum.
2. Electrical braking where the kinetic energy of the motor and tool is converted to electrical energy and is dispatching a heat in a resistance or returned to the supply system.

Electric braking is superior to mechanical braking since it is much quicker and eliminates the cost of maintenance of mechanical brakes. However, in order to finally bring the motor so a standstill and hold it there, friction brakes are essential.

The following types of electric braking are employed:

1. Plugging
2. Rheostatic or dynamic braking
3. Regenerative braking

PLUGGING

The connections of the armature are reversed so that the motor tends to rotate in the reverse direction thus providing the necessary braking effect. However, the supply must be cut off when the motor comes to rest otherwise it will start rotating in the reverse direction. Plugging may be employed with D.C. motor or induction motor and synchronous machines.

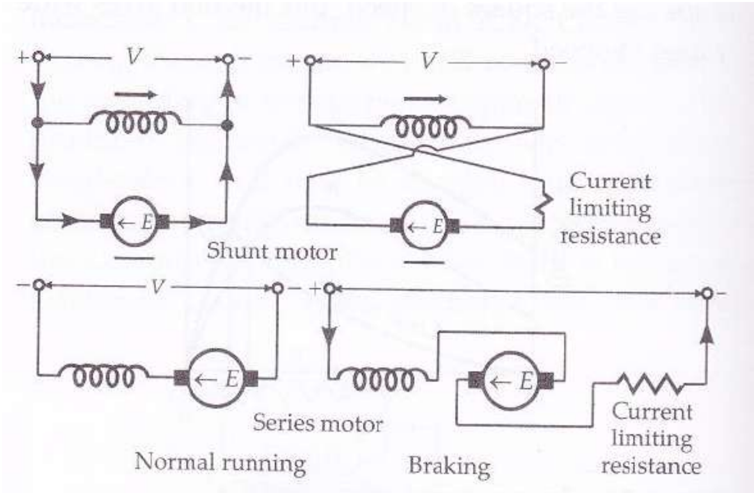


Fig.21 Plugging in DC motors

Plugging with D.C. motors

(a) The armature connections are reversed with respect to the field so that the current in the armature reverse .During normal running the back emf E is opposite to the direction of the armature current but during braking the back emf E and the armature current are in the same direction. At the instant of reversal of armature connections a voltage equal to V+E is impressed across the armature circuit, V being the supply voltage. Since E is very nearly equal to the V, the impressed voltage is approximately 2V.This will cause a great rush of current in the armature circuit. To prevent this, the starting resistance is reinserted in the armature circuit as shown below.

It should, however ,be noted that during braking, in addition to the kinetic energy of the motor being dissipated in the resistance, some energy is being drawn from the supply. There is, therefore, a waste of energy.

(b) If any two supply phases are interchanged with each other the direction of rotation of the magnetic field reverses and, therefore, the torque on the rotor also reverses providing a braking action. Supply, however, has to be cut off when the motor comes to rest, otherwise the rotor would start building up motion in reverse direction. The rotor and stator currents tend to be abnormally high and a resistance may have to be inserted in the rotor or stator circuit for the purpose of protection.

(c) PLUGGING WITH SYNCHRONOUS MOTORS

If the D.C. excitation of the synchronous motor is reversed, the D.C. and A.C. fields will rotate in opposite direction and there can be no braking effect. But in case of motors fitted within damper windings the eddy currents induced in them provide braking.

<u>Rheostatic or Dynamic Braking</u>

The motor is disconnected from the supply and worked as a generator driven by the kinetic energy of the rotor and the load. A resistance is connected across the motor terminals; the kinetic energy of rotation is converted into electrical energy and is dissipated in the resistance.

(a) D.C. Motors-shunt

The armature is disconnected from the supply and connected across a resistance. The motor now works as a separately excited generator and a braking torque is applied by the current delivered to the resistance. If, however, the supply fails, the braking action vanishes as the excitation disappears. This drawback is sometimes removed by fitting a series winding in the armature circuit which is connected during the braking period only. Due to the action of this winding, the motor self excites a series generator and the current delivered by the armature provides braking action.

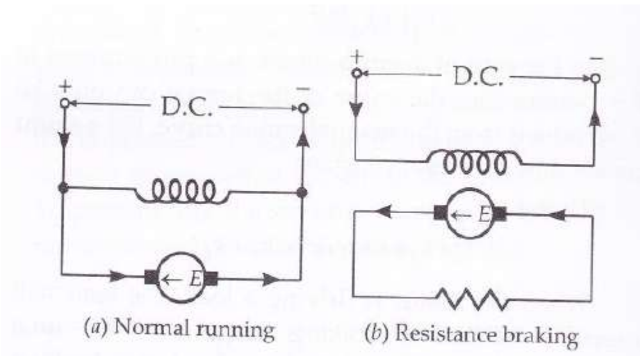


Fig..22 Rheostatic braking of shunt motor

(b) Series D.C. Motor- The motor after being disconnected from the supply is made to excite as a series generator. For this it is necessary that the total resistance in the motor circuit should be less than the critical resistance, so that the generator may self excite. Also in order that the flux may build up, the connections of the armature with respect to the field have to be reversed.

(c) Synchronous Motors

The field excitation is maintained and the motor after being disconnected from the supply is connected to resistances in star or delta. It now works as an alternator and the kinetic energy is dissipated in the form of losses in the resistances.

(d) Induction Motors

The stator is disconnected from the supply and direct steady current is passed through its windings .A flux is produced. When the short-circuited rotor conductors cut this steady flux emf is induced in them which provide the necessary braking effect. If the rotor is wound, the braking torque can be controlled by the insertion of suitable resistances in the rotor circuit.

Regenerative Braking

In regenerative breaking the motor is run as generator by the kinetic energy of the load which is returned to the mains as electrical energy. There is, therefore, an overall saving in energy.

(a) D.C.Motors-shunt: If the emf generated by the motor is greater than the supply voltage, power will be fed back into the supply. The emf in a shunt motor depends upon its excitation and speed. If the field is disconnected from the supply and the field current is increased by exciting it from another source, the induced emf will exceed the supply voltage and the motor will feed energy into the supply. The speed of the motor, however, falls to value corresponding to the field current at any instant. The condition is shown in Figure.

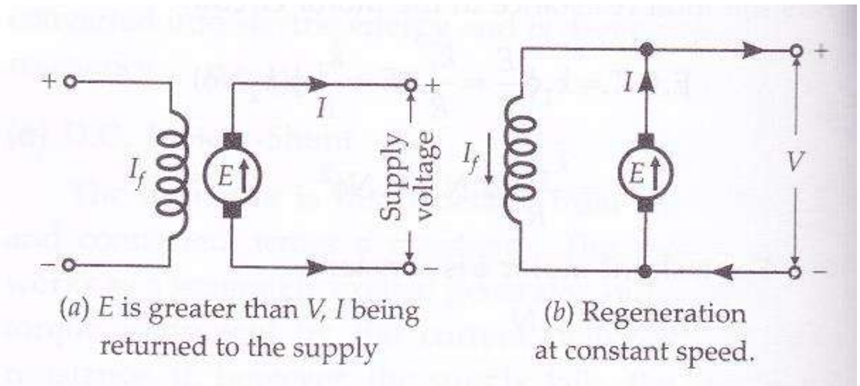


Fig.23. Regenerative Braking with d.c. shunt motor

There is another way in which regeneration takes place resulting in braking effect. If the field excitation does not change but the load causes the speed to increase, the induced emf may become greater than the supply voltage and power will be fed back in to the supply. The regenerative effect, however, will prevent any tendency of speed to increase further. This itself constitutes a form of braking effect since in the absence of regeneration, the speed would increase continuously.

(b) D.C. Motor-Series

Regenerative braking with series motors is employed mainly in traction work.

(c) Induction Motors

When an induction motor runs at a speed above the synchronous, it works as an induction generator feeding power back in to the supply. No extra devices need be employed .It may be noted that regenerative braking of induction motors is hardly useful for stopping the motor but it helps in keeping the load at a speed slightly above the synchronous and returns valuable power to the supply.

MECHANICAL FEATURES OF ELECTRIC MOTORS

The electric motor in this section is being discussed with reference to the following:

1. Type of Enclosure
2. Bearing
3. Transmission of drive

4. Noise

TYPE OF ENCLOSURE

This is mainly determined from a consideration of the type of work the motor has to do and the place where it is to be installed.

I. Open Type: The machine is open from both ends, the bearings being placed on pedestals or brackets. There is free ventilation. Such a type of machine can be used in a separate room.

II. Protected Type: Large operations are available for ventilation though some protection is provided. Mostly motors are of this type. If the openings are covered with a wire mesh or screen the motor becomes vermin proof and better protection is afforded without in any serious manner decreasing the ventilation.

III. Drip proof Type: Such motors are used where the atmosphere is very damp.

IV. Totally enclosed Type: Such type of motors is used where the atmosphere is dusty i.e. saw mills, coal-handling plants and stone crushing quarries. No foreign particle should enter the motor and block passage of ventilation. The ventilation facilities are very poor in view of closed construction. To improve this, the outer surface is finned to increase the cooking area. A totally enclosed motor fitted with an outer casing built round the motor proper through which clean air is circulated with a fan is common these days.

V. Pipe ventilated type: The large sizes of the totally-enclosed motor employ pipe ventilation. Clean air forced through the piping to cool the motor. The extra cost of piping is offset by the smaller size of motor that can be employed on account of better cooling.

VI. Flame-Proof Type: Necessary for explosive atmosphere met with in coal mines, chemical plants etc.

BEARINGS

The modern tendency is to use ball or roller bearings up to about 100 hp size as they have longer life and maintenance cost are low. These type of bearings are replacing old journal type bearings. They have enabled induction motors to be built with a very small air-gap. For large size motors and for reduction in the noise level, the journal type bearing is used.

TRANSMISSION OF DRIVE

The transmission of power from the driving machine may be arranged in various ways listed below. The choice of the motor speed is an important consideration. The cost per HP decreases with an increase in speed while the efficiency and power factor improve with speed. Therefore, a high speed motor is preferable to a low-speed one and if the speed of the work is to be low it can be achieved by reduction gears or other suitable means. Motors for low-speed drivers are manufactured now with the reduction gear incorporated in the unit itself so that the high speed motor gives a high efficiency and is less expensive.

Belt Drive: Maximum power of about 300hp can be transmitted. Slip is about 3 to 4 percent.

Rope Drive: A number of ropes are run in v-grooves over pulleys. It is used where the power is beyond the scope of belt drive. Slip is negligible.

Chain Drive: More expensive than the above two forms but more efficient and there is no slip. It can be used for greater speed ratio; the limit is 6:1.

Direct Drive: The motor is in the line with the driven machine. To protect the motor from sudden jerks, flexible couplings are used.

Vertical Drive: The motor shaft is vertical. The arrangement is sometimes convenient.

NOISE

In any industrial establishment, it is important to keep the noise level to as a low as value as possible so that fatigue to workers may be avoided. For domestic appliances, the motors used must be almost noise less and the same applies to appliances used in hospitals, \theatres, etc. The noise produced depends upon the loudness and shrillness of the note emitted. The level of sound is measured in decibels.

In motors noise may be produced due ti mechanical features like bearings, vibration and bad foundation. It is for the manufacturer to improve these. But the transmission of noise to other parts of a building cab is eliminated by foundations.

SIZE AND RATING

The size and rating of a motor for a given service dependent upon its temperature rise. The maximum temperature rise is limited by the type of insulation used. For class A Insulation a maximum temperature rise of 40^0C is permissible.For class B insulation value is 50^0C.

Standard Ratings for Motors: The I.S. specifications classify the motors for service as follows:

Continuous Rating: A motor capable of giving the output continuously without exceeding the rated temperature rise. It can also give 2.5% overload for 2 hours.

Continuous maximum Rating: Similar to the above without allowing any overload. It is used for motors of size larger than 2.5hp per rpm.

Short-time Rating: It is output that a motor can give for a specified short-time, say ½ hr. or ¼ hr. without exceeding the specified temperature rise.

MOTORS FOR PARTICULAR SERVICES

In the previous sections discussions were made regarding various aspects of performance of electric motors. The basic problem of choosing the proper motor is in matching the motor characteristics to the load requirements, i.e. load mechanics must be clearly known. On this depend the selection of the type of motor. After this comes the ability of the motor to carry the loads, this ability is limited by two factors; thermal and mechanical. Thermal consideration is with regard to the temperature rise under a given duty and mechanical is for ensuring that the peak load is carried by the motor safety.

let us consider one or two examples for choosing a motor. Suppose we want a motor for driving a fan. The load torque varies as the square of the speed. From among A.C. motors, induction motor is suitable. The stable point of operation is shown as S on the torque-speed characteristics in Figure.

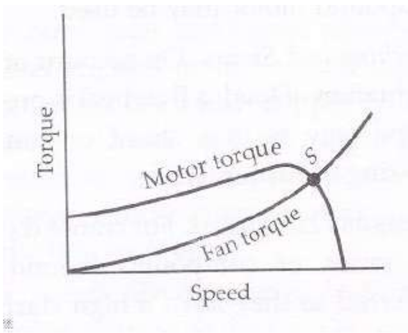


Fig.24 Fan load and induction motor torque speed curves

From D.C. motors shunt motor will be suitable .The point of stable operation is the point where the motor torque and the fan torque curves intersect each other as shown in figure.

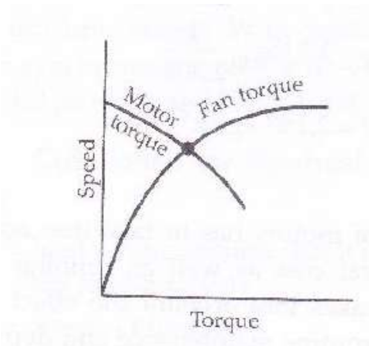


Fig.25. Fan load and shunt motor speed torque curves

Next let us consider a punching machine or shearing machine. The load is fluctuating in nature and therefore a flywheel will have to be used in order that the load demand as reflected on the supply may be smooth. A flywheel can only be used with a motor whose speed can decrease as the load comes. A D.C. shunt or compound or a 3-phase induction motor will be suitable.

Some of the motors commonly used for particular services are given below:

i) Domestic uses: Small universal motors of the series type are used in domestic appliances like vacuum cleaners, refrigerators, washing machine, fans etc.

ii) Grinding and milling machine: Upto 50 h.p. the motors may be D.C. shunt or induction with slip-rings and arrangement for pole-changing with cage rotors.

iii) Planners: There is cutting stroke and a quick return stroke. Arrangements for reversing the speed have to be incorporated. A D.C. compound motor may be used.

iv) Punching and shears: On account of heavy fluctuations of load, a flywheel is provided. The motor may be D.C. shunt or compound or slip-ring induction type.

v) Cranes and Hoist Work: For cranes D.C. motors of series or compound type are preferred as they have a high starting torque and the speed control is smooth. Induction motors are also used for hoisting.

vi) Lifts: duty involves high acceleration and high retardation. The motor armature must therefore be light and it should run at moderate speeds .D.C. compound, slip-ring induction, induction-repulsion motors are used.

vii) Textile Industry: Motors must be of totally enclosed type to prevent particles of the material being manufactured from getting into them. They should also be moisture proof on account of damp atmosphere inside a textile plant.

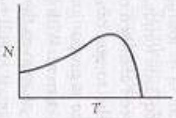
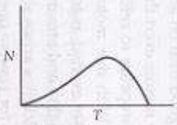
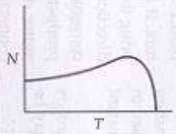
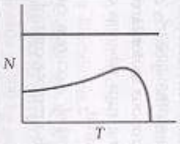
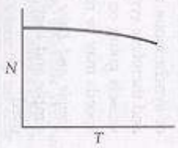
Three-phase motors are used since their speed is fixed by the supply frequency .D.C. motors cannot be used as their speed varies with voltage.

viii) Printing machinery: As it requires a variable speed Induction motors using rotor resistance may be used. Where large speed variations are required D.C. compound or A.C. commutator motors may be used.

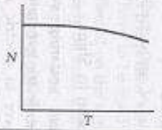
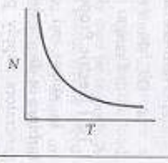
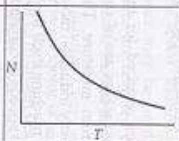
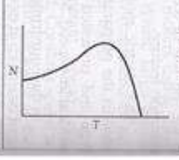
ix) Paper Industry: Synchronous motor is used as in this a constant speed of operation is required for giving a uniform thickness of the paper. Where speed is not required to be constant, squirrel cage induction motors or D.C. motors may be used.

x) Iron and steel Industry: D.C. shunt motors with flywheel arrangement or induction motors with speed control are used in such Industry.

xi) Mining work: Flame proof motors are required for safet6y considerations in a mines for which cage motors are used up to 10 hp and for large output slip-ring or D.C. motors are used.

Type of motor	Voltage limit	H.P. limit	$\frac{I_{starting}}{I_{full\ load}}$	$\frac{T_{starting}}{T_{full\ load}}$	Methods of	Speed-torque characteristic	Range of speed control	Methods of speed control	Applications
Squirrel-cage induction motor	11 V	300	1	1	• Start-delta starter • Auto-transformer starter. • Resistance in stator circuit.		Small variation in speed from no load value due to slip.	• Pole changing. • Variation of frequency.	Industrial crives of small power having explosive atmosphere. Water pumps and Tube-wells.
Single-phase induction motor	250 kV	1	Capacitor start 2 repulsion start 2	Capacitor start 1.5 repulsion start 3	• Repulsion start • Pole shading • Phase splitting by L or C and R in series with one of the windings.		Small variations due to slip.	By voltage drop in series impedance.	Fans, record-player-compressors, refrigerators, washing machines, medical, appartus, vacuum cleaners, air conditioning installations.
Pole-changing induction motor (Squirrel cage)	11 kV	300	6	1	• Direct switching up to 5 hp. • Start-delta starter. • Auto-transformer starter.		2 to 4 fixed ratios	• Pole-changing. • Variation of frequency.	Cranes, lathes drills, lifts-high speed winding is used for acceleration and low speed for raterdatio and landing.
Synchronous motor	15 kV	10000	2	3/4	• By pony motor • By damper windings.		Fixed speed	Fixed speed (which depend on frequency).	Motor generator sets, frequency chargers, fans, compressors, pumps, lineshafts, calenders and rolling mills. Small motors for clocks.
Discharge motor	1 kV	1000	2	2.6	• Applying low voltage. • Direct switching keeping the brushes in low speed position.		3 : 1	• Shifting the brushes. • Inserting impedance in secondary winding.	Paper making machines, printing presses, textile work, lifts, pumps, machine tools, belt conveyors.

Summary of the characteristic and the field of application of the various types of motors

Type of motor	Voltage limit	H.P. limit	$\frac{I_{starting}}{I_{full\ load}}$	$\frac{T_{startin}}{T_{full\ load}}$	Methods of	Speed-torque characteristic	Range of speed control	Methods of speed control	Applications
Shunt motor	3000 V	25000	2	2	Series resistance in armature removed in steps.		4 : 1	• Inserting variable resistance in field circuit. • Inserting variable resistance in armature circuit.	For driving line shafts, lathes, milling machines, conveyors, fans.
d.c. series motor	1500 V	3000	2	3	• Series resistance. • Series-parallel method in traction work.		3 : 1	• By field diverter. • By tapping the field. • Variable resistance in series with the motor. • Series-parallel control in traction work.	Traction ; haulage ; crans and moving heavy slides.
Compound motor	1500 V	3000	2	2.3	Series resistance in armature removed in steps.	Can be adjusted between that of pure shunt and pure series motor.	3 : 1	• Variable resistance in shunt field circuit. • Series field circuit. • Combination of 1 and 2.	Shears, punches, elevators, conveyors, heavy planners, rolling mills for intermittent high torque load.
d.c. series motor	500 V	3000	2	3	Variable voltage applied from the secondary of a transformer with tappings.		Full range with transformer	By voltage variation.	Traction work.
Ship-ring induction	11 kV	11000	2	2	Resistance in rotor circuit.		Small variation in speed from no load value due to slip.	• Variation of applied voltage. • Variation of frequency. • Pole-changing. • Cascading. • Rotor rheostat control • Injecting e.m.f. in rotor.	Generators, line shafts, lifes, pumps ; mills, winding machines, haulage.

CHAPTER-6

ELECTRIC TRACTION

Introduction

The system of traction involving the use of electricity is known as the electric traction .

In the earlier stages of the development of Electric traction two systems have been in use –D.C. at 1500 volts or 3000 volts and single-phase a.c. at 11 to 16kV using low frequency. The reasons for the adoption of low frequency rather than the standard 50-cycle frequency was that the series wound commutator was developed for satisfactory operation only up to about 25 cycles and the low frequency was suitable for the hydro-generators of the railways which had to generate their own power in the absence of any national grids that exist today. The d.c series motor has ideal characteristics for traction purpose. D.C. was already in use for tramways and in about 1905, on account of the better performance of d.c. series motor due to the introduction of the inter-poles and adoption of higher voltage with increased spacing of the substations the traction became economical. The two systems i.e. D.C. and A.C. developed and grew side by side.

In India we have the single-phase A.C. at 25kV,50 cycles is supplied to the locomotives which carries transformers and rectifiers. A.C. is converted into D.C. in the locomotive and traction motors are D.C. motors. However, recently A.C. traction motors are being attempted.

SYSTEM OF TRACTION

There are various systems of traction are commonly used such as

1. Direct steam engine drive
2. Direct internal combustion engine drive
3. Steam electric drive
4. Petrol electric traction
5. Battery electric drive
6. Electric Drive
7. Internal combustion engine electric drive

Direct Steam Engine Drive :

The steam engine drive used to be widely employed for railway work. In this drive the reciprocating steam engine is invariably used for getting the necessary motive power because of its inherent simplicity, operational dependability, and simplified maintenance, the simplicity of connections between the cylinders and driving wheels and easy speed control. It causes no interference to the communication lines running along the track. It is cheap for low density traffic areas and initial stages of communication by rail.

Direct Internal Combustion Engine Drive:

Direct internal combustion engine drive is widely employed for road transport. The efficiency of internal combustion engine at its normal speed is about 25 percent. It is self contained unit and it is not tied to any route. Initially the cost of vehicle and garage is very low. Speed control and braking system employed is very simple. It is cheap drive for the outer suburbs and country districts.

Steam Electric Drive:

A few locomotives employing steam turbine for driving a generator used for supplying current to electric motors have been built for experimental purposes.

Internal Combustion Engine Electric Drive :

In this drive the reduction gear and gear box are eliminated as the diesel engine is to drive the dc generator coupled to it at a constant speed. This type of drive has found considerable favour for railway work and locomotives of this type are becoming widely used.

Petrol Electric Traction:

This system has been used in heavy lorries and buses. Due to electric conversion it provides a very fine and continuous control which makes the vehicle capable of moving slowly at an imperceptible speed and creeping up the steepest slope without throttling the engine.

Battery Electric Drive:

In this drive the locomotive carries the secondary batteries which supply power to dc motor employed for driving the vehicle .Such a drive well suited for frequently operated service such as local delivery of goods in large towns with maximum daily run of 50 to 60 km, shunting and traction in industrial works and mines. The major limitation of this type of drive is the small capacity of the batteries and the necessity for frequent charging, speed range is also limited.

Electric Drive:

The drive of this type is mostly widely used. In this system of traction the vehicle draws electrical energy from the distribution system fed at suitable points from either a central power station or substations.

System of Electric Traction

Two types of vehicles are in use for electric traction. In one type they receive power from a distribution network while in the other type they generate their own power. The former type vehicles may use both a.c. or d.c. ; the latter type will be the diesel-electric car or train, petrol-electric truck, lorry and battery driven vehicles.

DC TRACTION MOTOR

Most suitable motors for dc system are the series and compound motors.

DC Series Motor:

The series motor used for traction purposes have following requirements

1. The dc series motor develops high torque at start which is essential for traction services.

2. The series motor is simple speed control method.

3. Power drawn from supply mains varies as the square root of the load torque.

4. Series motor is not suitable for regenerative braking as these are not electrically stable.

5. In case of dc series motor commutation is excellent up to twice full load so replacement of brushes in not required frequently.

6 In cases of dc series motors the flux varies as the armature current, torque corresponding to a given armature current, therefore is independent of line voltage.

7 In case of dc series motor up to magnetic saturation, torque developed is proportional to the square of the armature current. Thus dc series motor requires comparatively less increased power input with the increase in load torque.

8. The series motor when operated in parallel to drive a vehicle by means of different axles, share load almost equally even there is unequal wear of different driving wheels.

9. The dc series motor is simple and robust in construction.

AC TRACTION MOTOR:

AC Series Motor: Many single phase ac motors have been developed for traction purposes but only compensated series type commutator motor is best for traction. The construction of an ac series motor is similar to a dc series motor except that some modification such as whole magnetic circuit laminated, series field with as few turns as possible, large no of armature conductors, use of carbon brushes, numerous poles with lesser flux per pole. Compensating windings are provided to neutralize armature reaction and commutating or interpoles are provided for better performance in terms of higher efficiency and a greater output from a given size of armature core. The speed –Torque characteristics and the speed-current characteristics of compensated series type commutator motors are similar to those of D.C. series motor.Thea.c. Series motor is not suitable to suburban services where stops are frequent. It is being extensively employed on main line work on the continent and in America and provides good service.

If a d.c. series motor is worked on a.c. it would not operate in a satisfactory manner. Though the torque on the armature would be unidirectional, it would be at double the frequency since both the field current and the armature current reverse every half cycle.The alternating flux would cause heavy iron losses in the field and yoke.Heavy sparking would also take place at the brushes since the induced voltage and currents in the armature would be short-circuited at the time of commutation. The overall performance of the motor would be poor.

The difference between d.c. and a.c. operation can be understood by a reference to figure shown below.

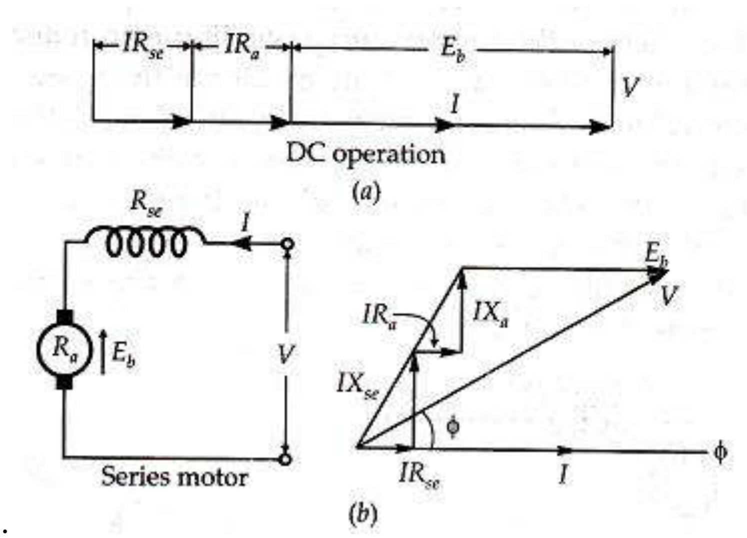


Fig. 1 Operation of series wound motor on dc and ac

Operation on d.c. is simple enough. I is the current drawn by the motor, IR_{se} and IR_a are the drops in the series field and the armature respectively. E_b is the back emf developed and equals kØN.

Mathematically, we have

$$V = E_b + IR_{se} + IR_a$$

Since $I(R_{se} + R_a)$ drop is about 10 percent of the applied voltage , E_b is practically equal to V.

On the a.c. the magnetizing component of the current and the flux are in time phase and the back emfE_b which is due to rotation of the armature is also in phase with the flux. If we neglect the loss component of the current we can assume the whole current to be in phase with the flux. The drops IR_{se} and IR_a are in phase with the current while the drops due to reactance, i.e. IX_{se} and IX_a are leading the current by 90^0. The a.c. operation is shown by the phasor diagram below. In this case E_b will be much less as compared to the d.c. operation. N is proportional to E_b and torque depends upon the product of E_b and I. Since, E_b in d.c. is larger than in a.c., for the same torque the speed for d.c. operation is higher than for a.c. operation as shown below.

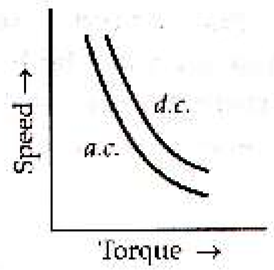


Fig.2 Speed – torque curves for d.c. and a.c. operation

In order to improve the performance of the motor on a.c., a compensating winding either in series with the armature or short-circuited in it be provided. The effect of the compensating winding is to reduce the armature reactance of the motor which increases the value of E_b and provides better speed regulation. The armature and field mmfs are at right angles to each other. The compensating winding provides an mmf opposite to the armature mmf and therefore considerably reduces the armature reactance drop. This is shown below

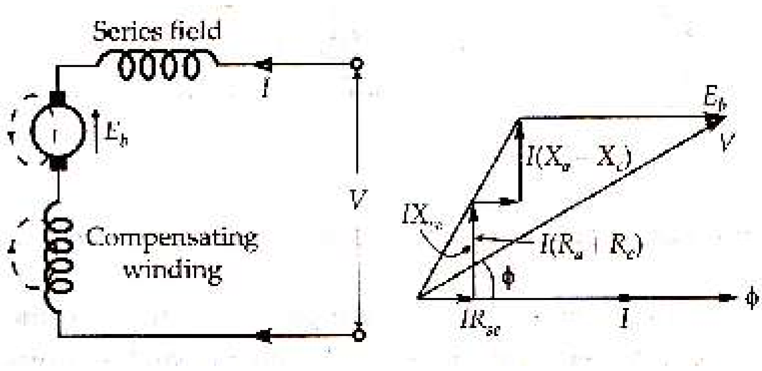


Fig. 3 Circuit diagram & phasor diagram of the series motor with compensating winding

$R_a + R_c$ represent the resistances of the armature and compensating winding.

$X_a + X_c$ represent the reactances of the armature and compensating winding.

Fig. below shows the case where the compensating winding is short-circuited on itself. It acts like the short-circuited secondary of a transformer and greatly reduces the effect of the armature reactance.In the phasor diagram R_a' and X_a' are the equivalent resistance and reactance of the armature and compensating winding referred to the armature circuit. It is also seen that by using the compensating winding ,the power factor of the motor improves as shown in the figure below.

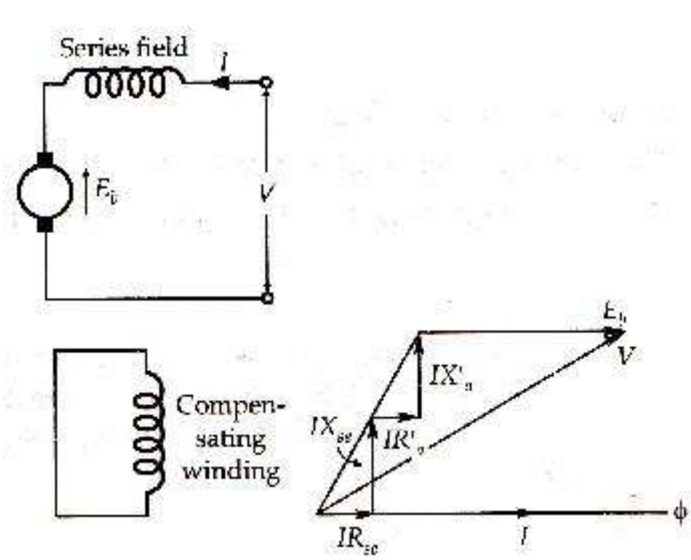


Fig.4. Circuit and phasor diagram for an inductively compensating series motor

THREE-PHASE INDUCTION MOTOR

Although it is robust and the simplest in construction, the difficulties in starting and speed control do not make it suitable for traction work. The speed torque curve is flat. It has been used in the Kando system in Hungary and some sections of Italian State Railways. It was not likely to find further application elsewhere though in recent years, with power electronic method of speed control, research is being undertaken to apply this drive in traction.

CONTROL OF MOTORS

CONTROL OF D.C. MOTORS

The starting current taken by a D.C. motor during its starting period is limited to a value approximately equal to the normal rated current by the resistance of the starter. There is a considerable loss of energy at the starting resistance. Consider the use of a single motor started by a resistance starter, the average value of the current during the starting period being limited to I, the normal full-load current.

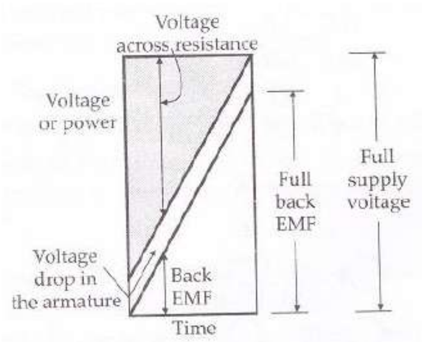


Fig.5. Voltage during the starting of a d.c. motor

The back emf of the motor starts to build up from zero magnitude. At the instant of switching on the supply,E_b=0, a current of I amperes is drawn from the supply and the supply voltage is the sum of the IR drop in the motor armature and the voltage drop across the starting resistance. At any other instant during starting, the supply voltage = (motor back emf)+(IR drop in the motor armature)+(voltage drop across starting resistance).

At the end of the accelerating period, the back emf has developed to a full value and the supply voltage =(back emf)+(IR drop).

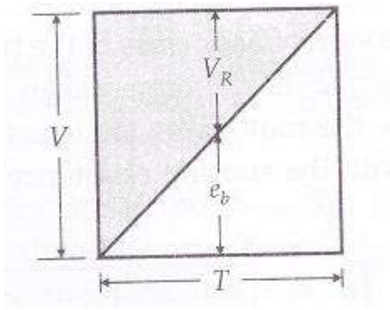


Fig.6. Starting of a dc motor by using a resistor in the armature circuit

If T is the time in seconds for starting and further if we ignore the voltage drop due to the resistance of the motor armature circuit we have total energy supplied =VIT watt-sec. in fig. shown the back emf and V_r is the voltage drop across the starting resistance st any instant.

SERIES PARALLEL STARTING

In traction work, usually two or more similar motors are employed. Considerable saving energy can be affected by employing series-parallel starting. Consider the use of two series motors. They are started in series with the help of a starting resistance till each of them develops a back emf equal to half the supply voltage minus the IR drop. The motors give one running speed when they are in the full series position. The starting resistance is again re-inserted in the circuit and the motors are switched in parallel. The starting resistance is cut out in steps and the back emf of each motor develops from about half the value to the normal value. In the full parallel position the motors give another running speed which is obviously higher than that when the motors are in full series.

Let us consider the case of two similar motors started by the series parallel method as shown below.

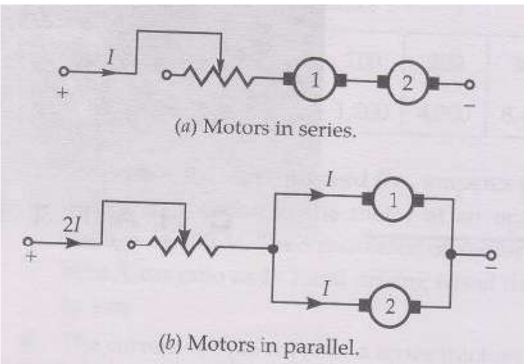


Fig.7. Series Parallel starting

Let the current during the starting interval be limited to the normal rated current I per motor. During the series period a current of I amperes is frawn from the supply while during the parallel period a current of 2I is drawn.

As shown in figure below, at the instant of starting OA =OB = IR drop in each motor,OK= supply voltage V. The back emf of the two motors jointly develops along the line OM .The back emf of two motors at the point E plus IR drops equal to V. Any point on the line BC at any instant represents the sum of back emfs of two motors and the IR drops.OE is the time taken for the series running.Now the motors are switched on in parallel, at instant E with the starting resistance reinserted.

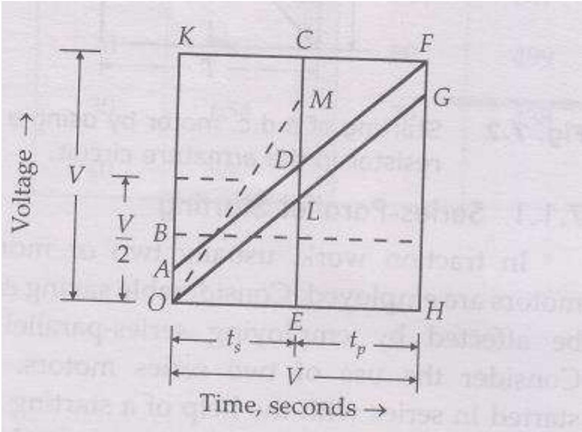


Fig.8 Voltage build-up in series-parallel starting

At the end of the series period each motor has developed a
Back-emf equal to (V- 2 IR)/2 = (V/2)-IR

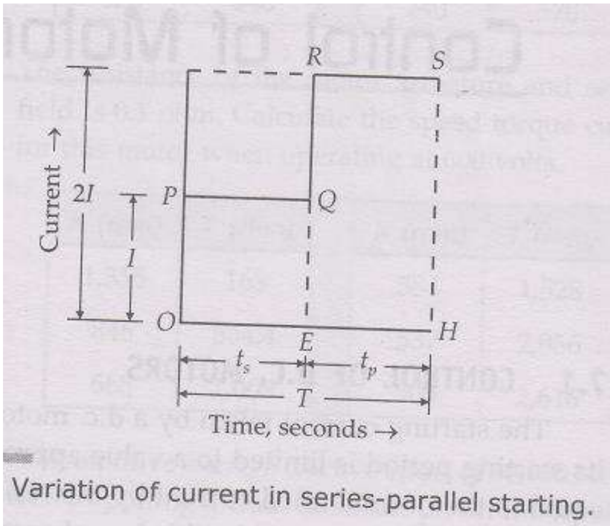


Fig.9 Variation of current in series –parallel

The back emf of each motor is represented by the ordinate EL = ED – LD =(((v/2)-IR)

The back emf of each motor is represented by the ordinate EL = ED-LD= ($\frac{V}{2}$ -IR)

The back emf of the combination now develops along LG and at H when the motors are in full parallel we have HF =supply voltage V,HG =normal back emf of each motor and GF =IR drop in each motor.

Figure below shows the current during the series and parallel starting periods.During the series period OE, the current is I while during the parallel period EH is 2I.

The value of the time t_s during which the motors remain in series and t_p , during which there are in parallel can be determined from figure shown below.Triangles OLE and OGH are similar.

Therefore

$$\frac{OE}{OH} = \frac{LE}{GH}$$

$$\frac{t_s}{T} = \frac{LE}{GH} = \frac{DE-DL}{FH-FG} = \frac{\frac{V}{2}-IR}{V-IR}$$

And $t_s = \frac{1}{2}\left(\frac{V-2IR}{V-IR}\right) T$

Hence t_p = T- t_s = T - $\frac{1}{2}\left(\frac{V-2IR}{V-IR}\right)$ T

= $\left[1-\frac{1}{2}\left(\frac{V-2IR}{V-IR}\right)\right]$ T

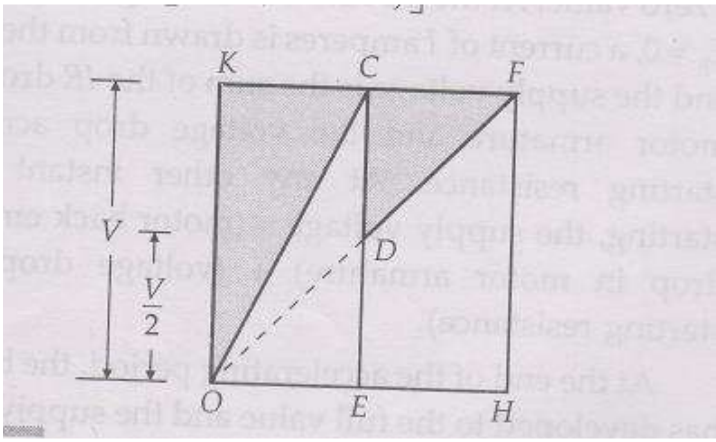


Fig.10 Efficiency of starting by series parallel method

Let us now calculate the efficiency of this method. For this purpose neglect the IR drop in the armature circuit as back emf developed practically equals the voltage impressed across the motor. This modifies the figure to as shown.

Since, D is the mid-point of CE and the back emf of the motor develops along DF in the parallel combination, KC = CF, i.e., time for series combination = the time for the parallel combination.

Let $t_s = t_p$ =t and the average starting current be I per motor, t_s =OE, t_p =EH.

The energy lost in starting resistance is proportional to the shading area. i.e.

= I($\frac{1}{2}$Vt) +($\frac{1}{2}\frac{V}{2}$ t) 2I =IVt

Total energy supplied $= IVt + 2IVt = 3IVt$

Efficiency of starting $= \frac{(3-I)IVt}{3IVt} = 2/3$ or 66%

Thus the efficiency is increased by about 17 % .The series-parallel method enables a saving of about 15 to 20 % in the energy.

The Series-Parallel Control

The series-parallel control is carried out as follows:

(a) Shunt Transition: The various stages involved in this method of series-parallel control are shown below.

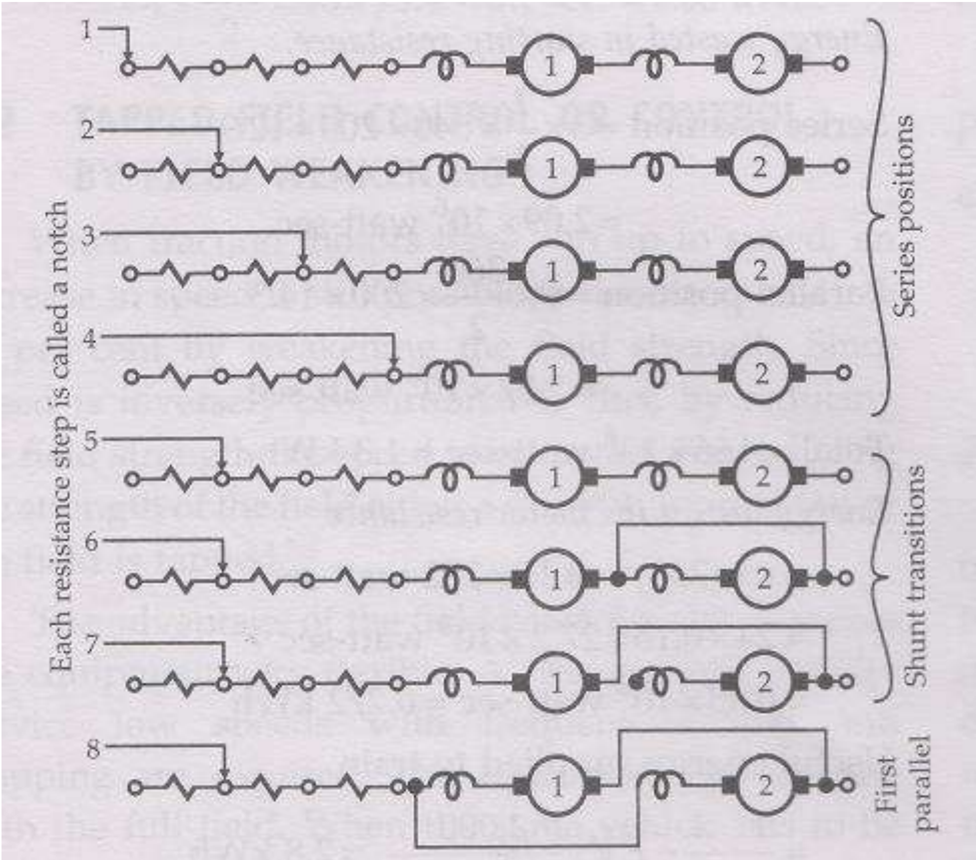


Fig..11 Series position

In steps 1,2,3,4 the motors are in series and are accelerated by cutting out the starting resistance in steps. In step 4 , the motors are in full series. During transition from series to parallel, the resistance is re-inserted in the motor circuit (step-5).One of the motors is by-passed (step-6) and disconnected from the main circuit (step-7).

It is then connected in parallel with the other motor (step-8) giving the first parallel position. The resistance is then cut out in steps completely and the motors are placed in parallel.

This method is known as the shunt-transition method.

(b) Bridge Transition: the motor and the starting rheostats are connected in the form of a Wheatstone bridge as shown below.

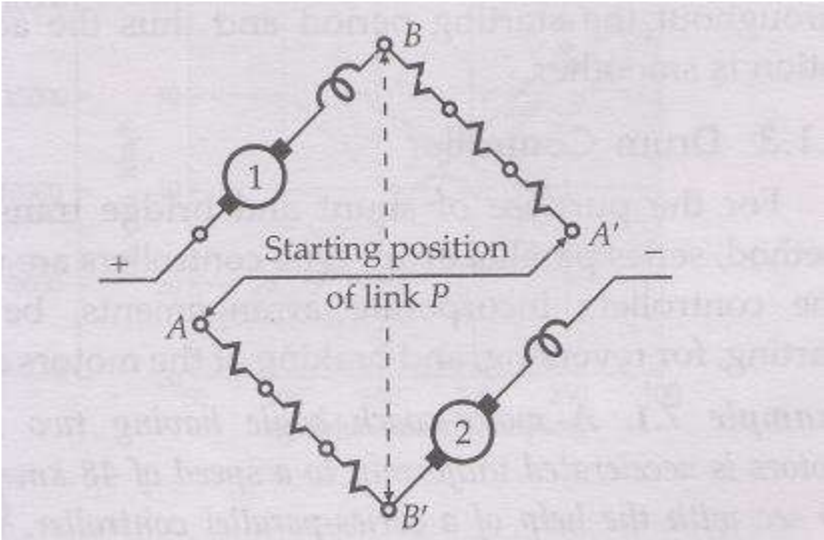


Fig.9. Series position.

(a) At starting, motors are in series with link P in position AA'

(b) Motors in full series with link P in position BB'

In the first starting position the motors are in series and the rheostats are completely in

Circuit as indicated by the rheostats arm P at A A′. A and A′ are moved in the direction of the arrowheads and in position BB′ the motors are in full series.

In the transition step, the rheostats are reinserted by connecting to positive and negative of the supply as shown below.

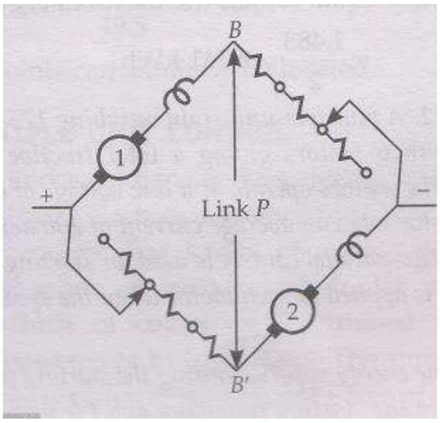


Fig.10 Transition step

In the first parallel step, the link P is removed and the motors are connected in parallel with the starting resistances in their circuit.

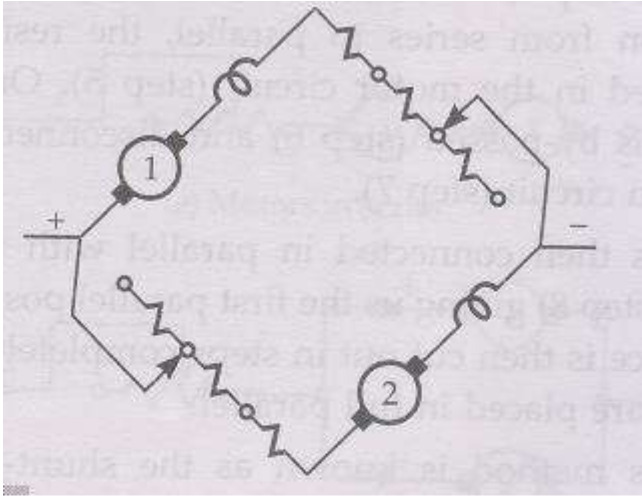


Fig.11 First Parallel position

The advantage of the bridge transition method over the shunt transition methid is that the normal accelerating torque is available from both the motors throughout the starting period and thus the acceleration is smoother.

DRUM CONTROLLER

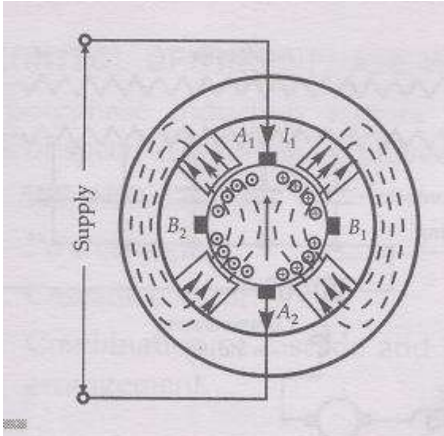
For the purpose of shunt and bridge transition method, series parallel drum type controllers are used. The controllers incorporate arrangements, besides starting, for reversing, and braking of the motors also.

Tapped Field Control :As the speed of the motor is inversely proportional to the flux (assuming line voltage constant), therefore, the speed can be varied by varying the flux. In case of series motors the flux can be varied either (i) by connecting a variable resistance known as diverter in parallel with the series field winding or (ii) by cutting out some of the series field turns. Since in both the cases the flux can be only reduced, therefore, this method is known as field weakening method and speeds above normal can be obtained. By this method speed can be raised to the extent of 15 to 30 percent of normal speed owing to design difficulties arising with traction motors.

The field weakening method is of no use for starting purpose. This method is used for increasing the speed of traction motors upto the extent of 10 to 15 percent when they have attained maximum possible sped by series-parallel control system. The advantage of this system is that it increases the flexibility of the train utility.

THE METADYNE SYSTEM OF CONTROL FOR D.C. MOTORS

In the series-parallel control of D.C. traction motors,there is considerable loss of energy in the starting resistances. The metadyne system of control estimates the energy loss and achieves a very smooth control during the acceleration period.



Illustrating Metadyne.

Fig.12 Illustrating Metadyne

Consider a D.C. armature with two brushes and two poles. If current is supplied to the two brushes A_1A_2 the armature cross-flux will be as shown anmd mainly confined to the poles as shown in Figure. If there are four brushes, current is supplied to brushes A_1A_2 and the armature cross-flux will take up the path as shown below.If now the current is supplied to brushes B_1B_2 as shown the armature cross-flux takes up path as indicated.If the armature is rotated at a constant speed and a current I is fed into the bushes A_1A_2 , an emf is induced in the winding betweenB_1B_2 due to the flux produced by I.No emf is induced between A_1A_2 and the voltage between A_1A_2 is on account of the voltage drop due to I_1 .

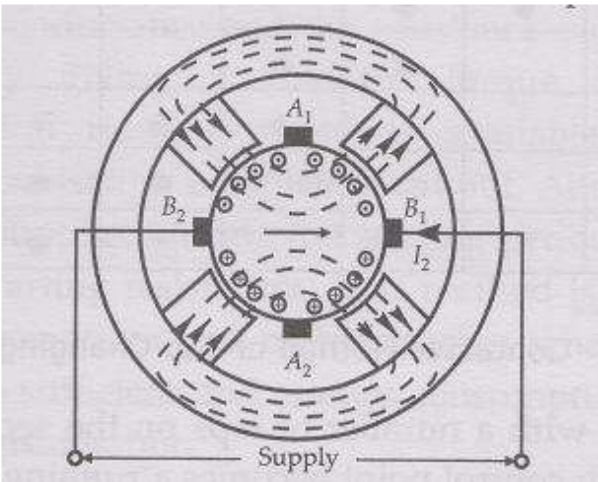


Fig.6.14

Since an emf is induced across B_1 ,B_2 a current I_2 will flow in a load connected between them. The resultant flux distribution on account of I_1 and I_2 is as shown below.

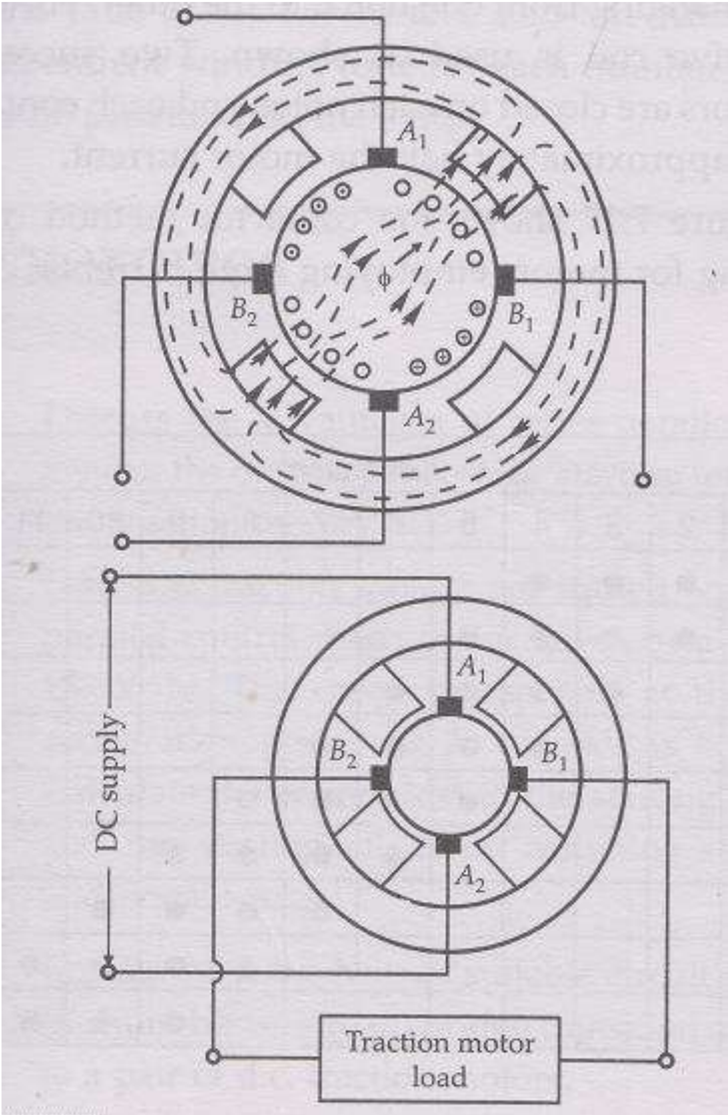


Fig.15

The total flux may be assumed to be made up of two components $\emptyset_1$ and $\emptyset_2$ at right angles and directed along A_2A_1 and B_2B_1 .The rotation of the armature in $\emptyset_2$ induces an emf E_1 between A_1 and A_2 which opposes the supply voltage.Since the current is to be kept at its original value of I_1,the supply voltage must be induced to overcome E_2 . Under steady conditions

$E_1 \propto \emptyset_2 = KI_2$

$E_2 \propto \emptyset_1 = KI_1$

$E_1I_1 = E_2I_2 = K\ I_1I_2$

This shows that the machine behaves like a D.C. transformer. Only the rotational losses of the machine need be supplied by the driving motor.

If the supply voltage E_1 remains constant , I_2 remains constant. The arrangement therefore is quite suitable for starting D.C.motors

RheostaticControl :A series motor can be started by connecting an external resistance (starter) in series with the main circuit of the motor. At the starting instant, since the back emf developed by the motor is zero, therefore, the resistance connected in series with the motor is maximum and is of such a value that the voltage drop across it with full load rated current is equal to the line voltage. As the motor speeds up, the back emf developed by the motor increases, therefore, the external resistance is gradually reduced in order to maintain the current constant throughout the starting or accelerating period. Basic traction motor circuit with rheostatic starting is shown in figure. In this method there is a considerable loss of energy in the external circuit.

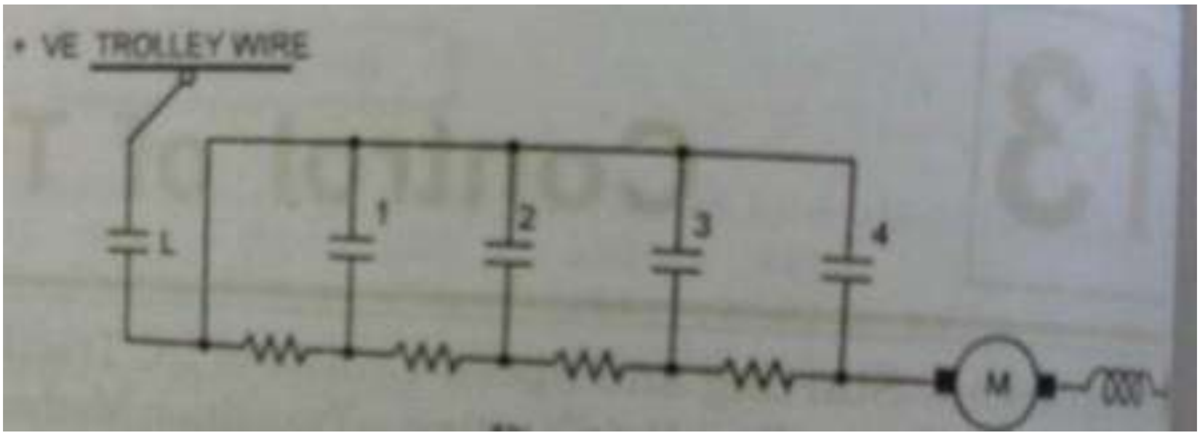


Fig.16 Rheostatic control method

BRAKING

Introduction

In traction work both electrical and mechanical braking are employed for bringing the vehicle to rest. Electrical braking cannot do away with the mechanical brakes since a vehicle cannot be held stationary by its use; it nevertheless forms a very important part of a traction system..The main advantage of using electric braking is that it reduces the wear on the mechanical brakes and gives a higher value of braking retardation thus bringing a avehiclequickly to rest and cutting down considerably on the running time.Where regenerative braking is employed , a part of the energy is returned to the supply thereby affecting a considerable saving in the running costs.

For D.C. motorsThere are three methods emp-loyed for electric braking:

(i) Plugging
(ii) Rheostatic braking
(iii)Regenerative braking

Plugging : Elaborate discussions have already been made on this in previous chapter and does need any more of it.

Rheostatic Braking

When two or three series motors are used for traction work,the motors are connected in parallel across a resistance.The kinetic energy of the vehicle is utilised in driving the motors as generators which dissipate this energy in the form of heat in the rheostats to which they are connected. The two machines in parallel amount to two series generators in parallel and in order that they may self-excite, an equalizer connection as shown has to be used. If the equalizer connection are not used,the machine that would build up first would send a current through the other in the opposite direction with the result that the second machine would excite withreversedvoltage.The two machiones would be short-circuited on themselves and might even burn out on account of large currents. The equalizer prevents such a condition.

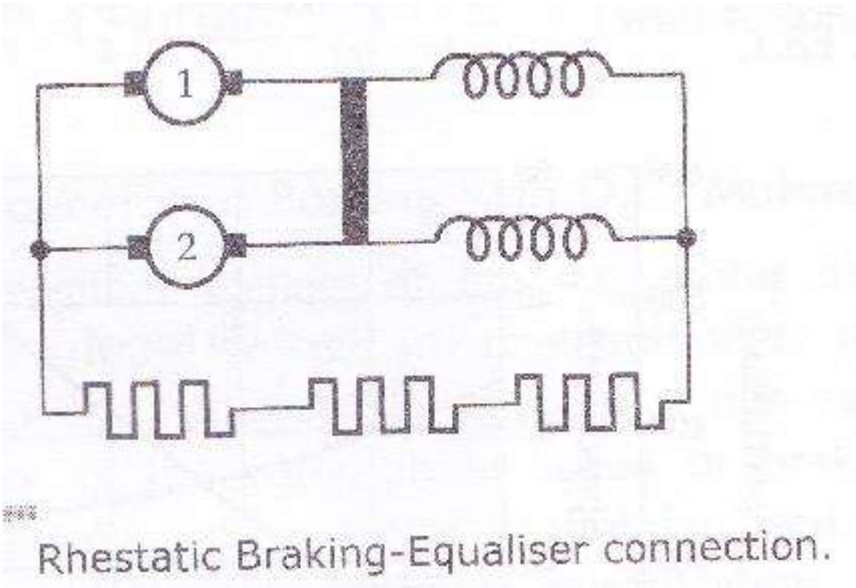


Fig.17

Another way to cross connect the fields of the machines is shown below.

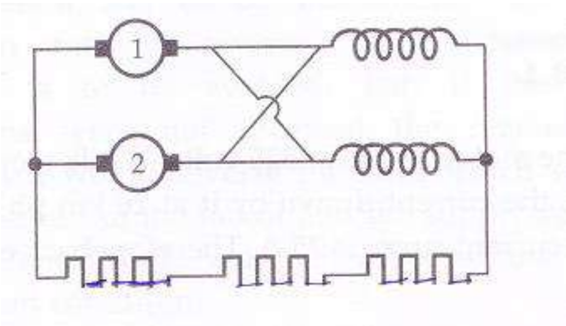


Fig.17 Rheostatic braking cross connection

Suppose the voltage of machine 1 is greater than that of 2.It will send a greater current through the field of machine 2 causing it to excite to a higher voltage and its own excitation will be kept down because of the lesser induced emf of 2.Thus automatic compensation is provide and the two machines operate satisfactorily.

The connections of the second case possess an advantage over that of the first. If the direction of rotation of the armature reverses, say, due to run-back, the machines fail to excite in the first case and no braking effect can be produced. However, with the cross-connected fields the machines build up in series and since they are short-circuited upon themselves, they provide emergency braking and would not allow the car to run-back on a gradient.

REGENERATIVE BRAKING

Mechanical Regenerative braking

When a train is accelerated up to a certain speed, it acquires kinetic energy corresponding to that speed. During the coasting period, a part of this kinetic energy is used up in overcoming the fractional resistance and some part is utilized in the propulsion of the train. The kinetic energy ,which is utilized in the propulsion ,does useful work and therefore coasting may be regarded as " mechanical regenerative braking " since the speed gradually decreases on account of the utilization of the kinetic energy stored in the train at the end of the accelerating period.

A system of track grading is employed in the case of the underground railway where the kinetic energy of a train may be used in doing useful work against gravity. Two types of graded tracks are shown in Figure below.

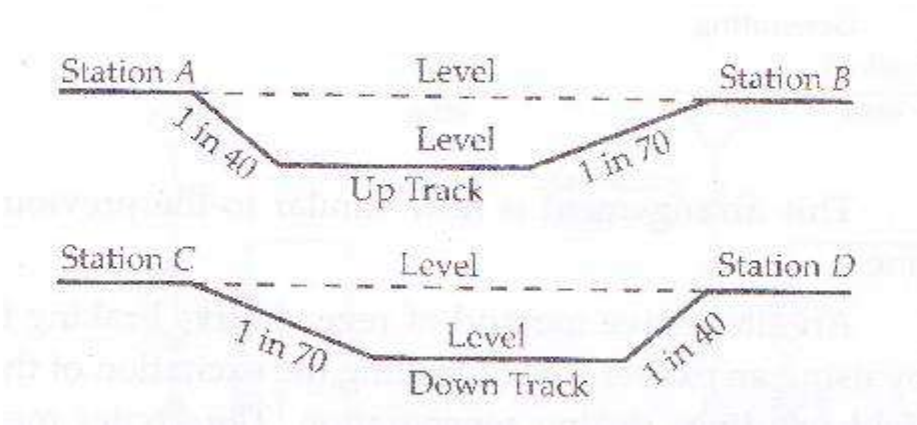


Fig.18 Track Grading

When the train is at a station, say A, it possess a certain potential energy which is utilized in its descent down the gradient till it reaches the level tracks. However, graded track construction is only possible in the case of the underground railway and is not practicable for surface railway.

Regenerative Braking with D.C. Motors

The terminal voltage of the D.C. motor must exceed the supply voltage for regeneration to take place. Also this voltage must kept at this value irrespective of the variation in speed or braking torque. The D.C. series motor cannot be used for regenerative braking without modification for the reasons to be stated presently. During regeneration the current through the armature reverses and since the excitation has to be maintained, the field connection must be reversed, if a short-circuit condition to be avoided. For, if the field connection were not reversed the regenerated current in it would reverse the field which would reverse the emf of the motor and the supply voltage and back emf would aid each other setting up a short-circuit condition.

One method of regenerative braking with series motors is the French method. If there is a single series motor as in the case of a trolley-bus it is equipped with a main series field auxiliary any field windings placed in parallel with the main series winding.

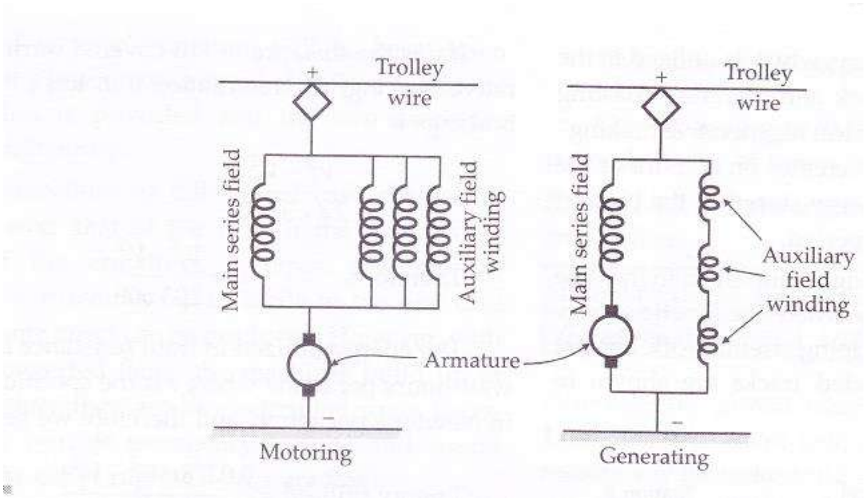


Fig. 20 Regenerative braking

During regenerative breaking the auxiliary field windings are placed in series with each other and switched over in parallel across the armature and the main series field. The machine acts as a compound generator with slight differential compounding. If there is a change in the line voltage, the shunt excitation being sensitive to such changes, immediately causes the emf of the generator to increase or decrease thus providing the necessary balance. Suppose the line voltage tends to increase beyond the emf of the generator. The increased voltage across the shunt circuit will send a large exciting current through it causing the emf of the generator to rise. The reverse of this happens when the line voltage tends to fall. The arrangement is, therefore, self-compensating.

In locomotives where four or six series motors are used, there need not be any auxiliary windings. During normal working all the motors are in series within their respective field windings but during regeneration, the motor armature is in parallel with the field windings of all other motors except one.

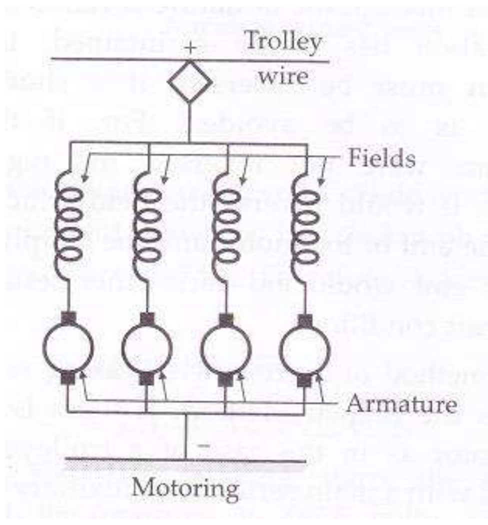


Fig.21 Regenerative Braking

This arrangement is now similar to the previous one. An alternative method of regenerative braking by using an exciter for controlling the excitation of the field winding during regeneration. The exciter may either be axle driven or noticed from an auxiliary supply.

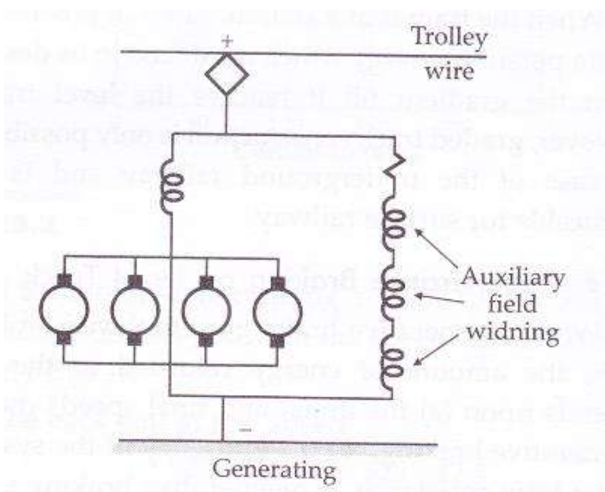


Fig. 22 Regenerative braking (alternative)

As shown in Figure the exciter E have separately excited winding whose excitation is controlled by the driver. The armature of the exciter is placed in the circuit of the series fields of motors 1 and 2. The exciter has other winding F_2 placed in series with the main motor circuit.F_2 and F_2 are arranged to oppose each other during regeneration.Suppose the line voltage decreases, it will try to increase the regeneent through the armayures 1 and 2.The excitation of F_2 therefore increase and since F_2 and F_2 each other, the emf of E falls on account of reduced excitation. As soon as the emf of the exciter falls , the current in the field current 1 and 2 decreases causing the emf of 1 and 2 to decrease. Compensation for a decrease in the line voltage is automatically provided, The arrangements shown below has the exciter connected in the circuit of the field windings and the stabilizing resistance. The balance of voltage available in the exciter armature circuit is reduced causing a reduction in the exciting current in the fields of 1 and 2.this decrease the induced emf of the generators, thus providing inherent compensation.

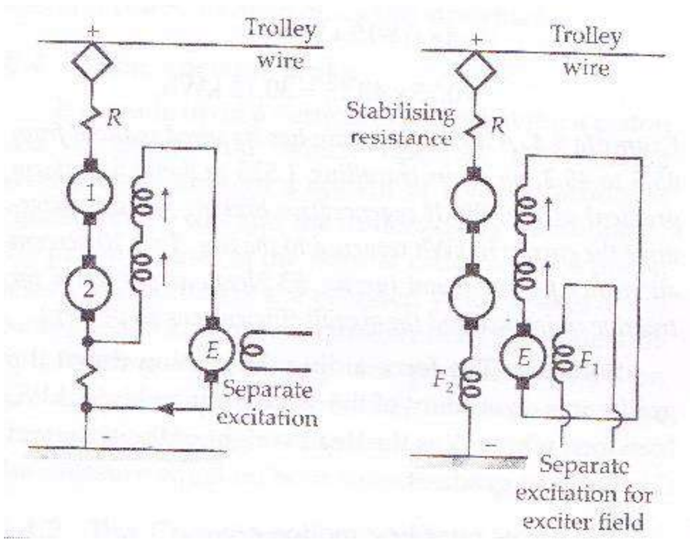


Fig. 23 Regenerative braking

The function of the stabilizing resistance is to prevent the current surges when the vehicle crosses one section of the supply to another and to compensate for variation in line voltage.

Limits of Braking

Regenerative braking is employed down to a speed of 16 km.ph. Then rheostat braking to about 6.5 km.ph and then mechanical brakes are used to bring the vehicle to rest.

REGENERATIVE BRAKING WITH THREE-PHASE INDUCTION MOTORS

Regenerative braking with three-phase induction motor occurs automatically when the motor runs at a speed slightly above the synchronous. It then works as induction generator. The induction generator however is not self-starting and must be connected to a system supplied from synchronous generator.

The torque-curve of an induction motor is as shown below. With no extra resistance in the rotor circuit, there is only a slight variation of speed with torque. By adding extra resistance in the rotor circuit the speed increases for a particular braking torque.

Therefore while braking without any extra resistance in the rotor circuit; the speed will be kept almost constant independent of the gradient and the load of the train. This is a great advantage with the induction motor when used for traction.

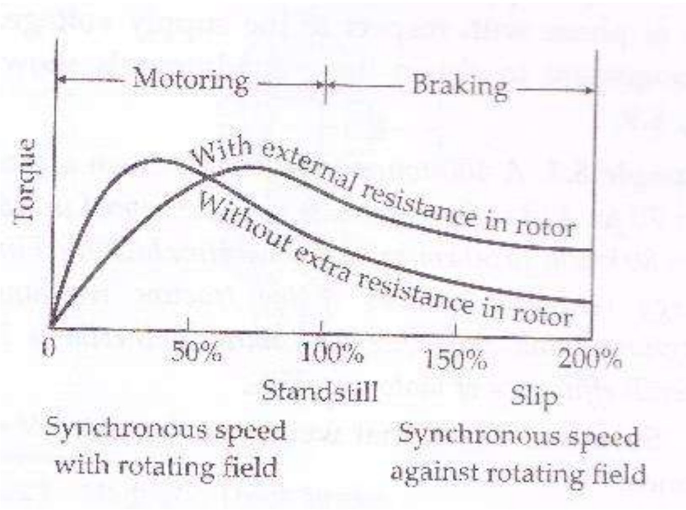


Fig.24 Torque speed curve of an Induction motor

BRAKING WITH SINGLE-PHASE SERIES MOTORS

In this case both rheostatic and regenerative braking are possible.

Rheostatic Braking: The motors are worked as separately excited generators supplying energy to resistance load. The fields are energized at low voltage from suitable tapings on the train transformer. The kinetic energy of the rotor is dissipated as electrical energy in the load resistance. Also, the fields of the motors may be excited from one of the motors acting as a series generator. In this case D.C. will be generated in the rotors of the motors and the kinetic energy of rotors will be dissipated as D.C. power in the loading resistors.

Regenerative Braking

For generative braking the regenerated power should be at the frequency of the main supply. This necessitates the energizing of the field winding from the main supply. Secondly, the regenerated current must be in phase opposition to the applied voltage and also the flux Ø so that the power may be feedback into the supply system.The voltage applied to the field winding must be 90^0 out of phase with respect to the supply voltage.An arrangement to obtain these conditions is shown below.

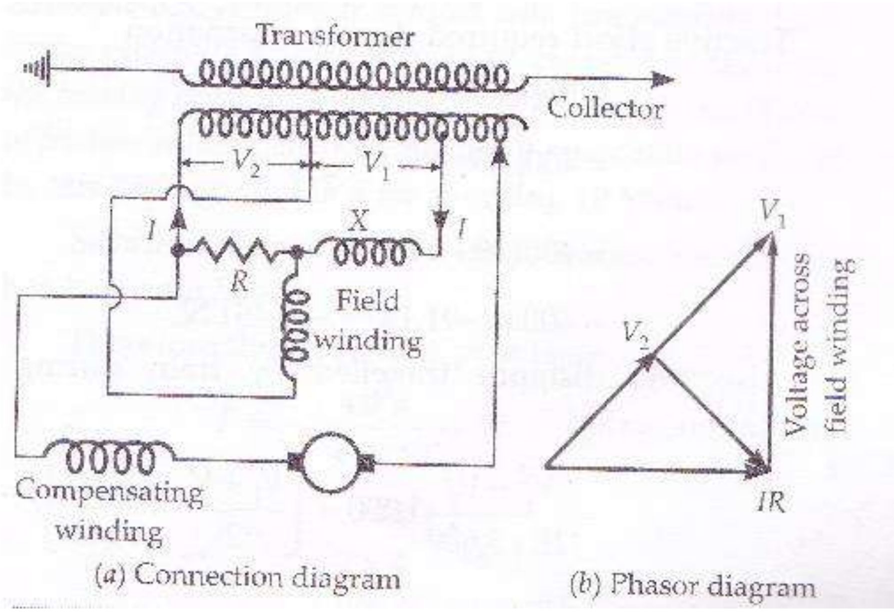


Fig.25 Regenerative braking with single phase series motor

MECHANICAL BRAKING

Mechanical brakes are essential feature on traction vehicles and are always operated by power. Two types of mechanical power brakes have been developed. (i) compresses air-brakes and (ii) vacuum brakes. The compressed air brake is extensively used on electrified railway and vacuum brakes on steam railway. The compressed air brake possess a little advantage over the other type since compressed air can conveniently be stored up and released for quick action where as the vacuum brake, a pump has to create the necessary vacuum. However, use of a vacuum reservoir overcomes this drawback.

THE VACUUM BRAKE

It is made up of a vertical cylinder having a piston and a piston rod which operates the braking arrangement through a system of levers. Vacuum is created on the top and the underside by admitting of the piston so that in the normal condition, the piston rests at the bottom of the cylinder. When brakes are to be applied, the vacuum is broken from the underside by admitting air at atmospheric pressure. The piston moves up applies the brakes. The brakes may be released by either creating the vacuum or by making the pressure equal on both sides of the piston.

The Compressed Air Brake

It consists of a reservoir of compressed air, a brake cylinder, a valve and pipe. The brakes are kept in the: off” position by springs in the brake cylinder. When brakes are to be applied, compressed air is admitted into the cylinder. It presses the piston against the force of the spring. Clearly, the force with which the brakes are applied depends upon the quantity of the compressed air admitted. To release the brakes, compressed air is exhausted from the cylinder.

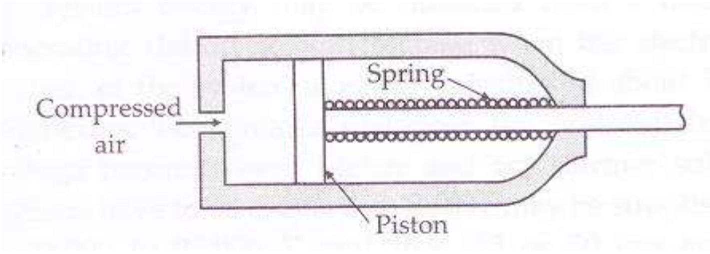


Fig.26 Action of compressed air brake

Magnetic Track Brake

It is used in tramcars. The electromagnet is bipolar. The body is made of cast steel and the pole faces are made of soft steel and can be renewed. The exciting coil is enclosed in a water-tight case. The magnetic flux is perpendicular to the pole faces and the track. The force of attraction between the magnet and the track is given by

$F = \frac{B^2 a}{2\times\pi\times10^{-7}}$N, where B is the flux density in weber/m^2 and a is the area in the pole face in sq.m.The drag that it can produce on the car is given by micro farad, where t is the coefficient of friction.

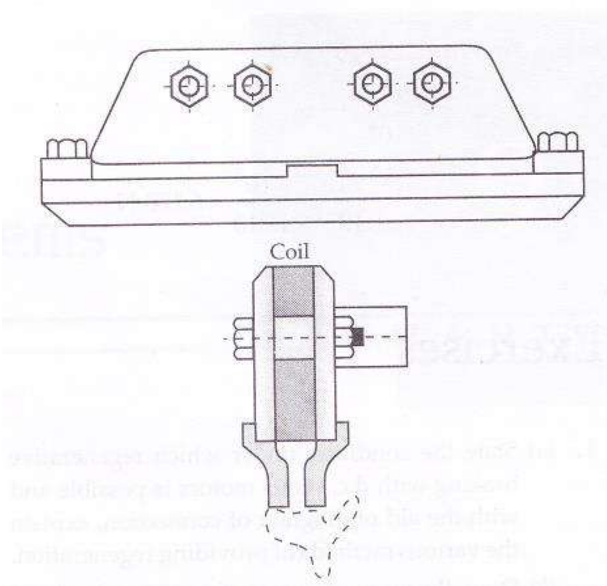


Fig.27 Magnetic Track brake

Electro-Mechanical Drum Brakes

The brake drum is fitted to the motor shaft and brake shoes are applied by springs and released by a solenoid excited from a battery. They have replaced the hand applied wheel brake.

GANDHI ACADEMY OF TECHNOLOGY AND ENGINEERING

GOLANTHARA, BERHAMPUR

LECTURE NOTES

ON

CONTROL SYSTEMS AND COMPONENT

For 6th Semester

ELECTRONICS AND TELECOMMUNICATION

(As per Syllabus prescribed by SCTE&VT, Odisha)

Prepared By

Mr. Susanta Kumar Sahu

(Lecturer in Electrical Engineering)

CHAPTER-01

FUNDAMENTAL OF CONTROL SYSTEM

A **control system** manages commands, directs or regulates the behavior of other devices or systems using control loops. It can range from a single home heating controller using a thermostat controlling a domestic boiler to large Industrial control systems which are used for controlling processes or machines. A control system is a system, which provides the desired response by controlling the output. The following figure shows the simple block diagram of a control system.

Examples − Traffic lights control system, washing machine

Traffic lights control system is an example of control system. Here, a sequence of input signal is applied to this control system and the output is one of the three lights that will be on for some duration of time. During this time, the other two lights will be off. Based on the traffic study at a particular junction, the on and off times of the lights can be determined. Accordingly, the input signal controls the output. So, the traffic lights control system operates on time basis.

Classification of Control Systems

Based on some parameters, we can classify the control systems into the following ways.

Continuous time and Discrete-time Control Systems

- Control Systems can be classified as continuous time control systems and discrete time control systems based on the **type of the signal** used.
- In **continuous time** control systems, all the signals are continuous in time. But, in **discrete time** control systems, there exists one or more discrete time signals.

SISO and MIMO Control Systems

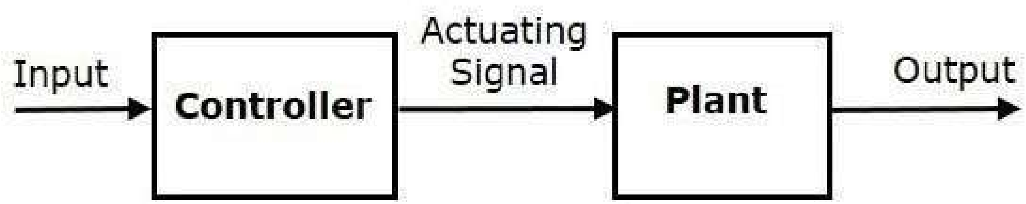
- Control Systems can be classified as SISO control systems and MIMO control systems based on the **number of inputs and outputs** present.
- **SISO** (Single Input and Single Output) control systems have one input and one output. Whereas, **MIMO** (Multiple Inputs and Multiple Outputs) control systems have more than one input and more than one output.

Open Loop and Closed Loop Control Systems

Control Systems can be classified as open loop control systems and closed loop control systems based on the **feedback path**.

In **open loop control systems**, output is not fed-back to the input. So, the control action is independent of the desired output.

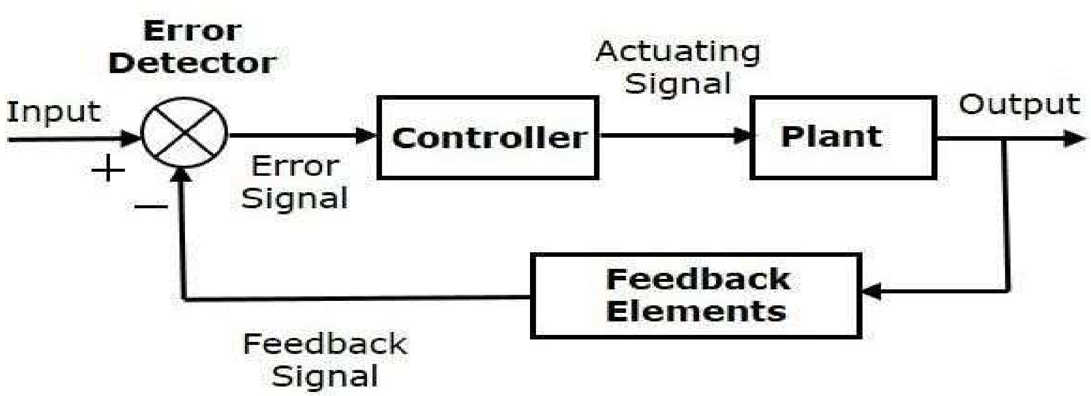
The following figure shows the block diagram of the open loop control system.



Here, an input is applied to a controller and it produces an actuating signal or controlling signal. This signal is given as an input to a plant or process which is to be controlled. So, the plant produces an output, which is controlled. The traffic lights control system which we discussed earlier is an example of an open loop control system.

In **closed loop control systems**, output is fed back to the input. So, the control action is dependent on the desired output.

The following figure shows the block diagram of negative feedback closed loop control system.



The error detector produces an error signal, which is the difference between the input and the feedback signal. This feedback signal is obtained from the block (feedback elements) by considering the output of the overall system as an input to this block. Instead of the direct input, the error signal is applied as an input to a controller.

So, the controller produces an actuating signal which controls the plant. In this combination, the output of the control system is adjusted automatically till we get the desired response. Hence, the closed loop control systems are also called the automatic control systems. Traffic lights control system having sensor at the input is an example of a closed loop control system.

The differences between the open loop and the closed loop control systems are mentioned in the following table.

Open Loop Control Systems	Closed Loop Control Systems
Control action is independent of the desired output.	Control action is dependent of the desired output.
Feedback path is not present.	Feedback path is present.
These are also called as **non-feedback control systems**.	These are also called as **feedback control systems**.
Easy to design.	Difficult to design.
These are economical.	These are costlier.
Inaccurate.	Accurate.

If either the output or some part of the output is returned to the input side and utilized as part of the system input, then it is known as **feedback**. Feedback plays an important role in order to improve the performance of the control systems. In this chapter, let us discuss the types of feedback & effects of feedback.

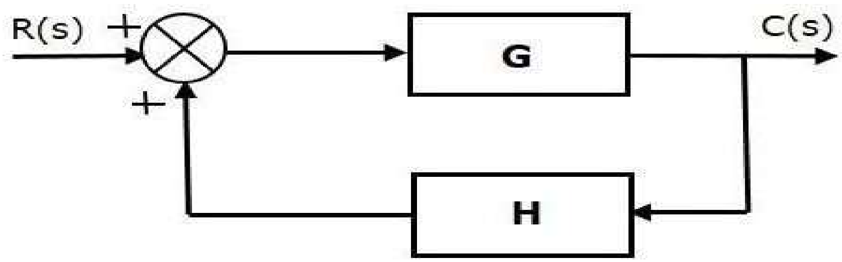
Types of Feedback

There are two types of feedback −

- Positive feedback
- Negative feedback

Positive Feedback

The positive feedback adds the reference input, R(s)R(s) and feedback output. The following figure shows the block diagram of **positive feedback control system**



he concept of transfer function will be discussed in later chapters. For the time being, consider the transfer function of positive feedback control system is,

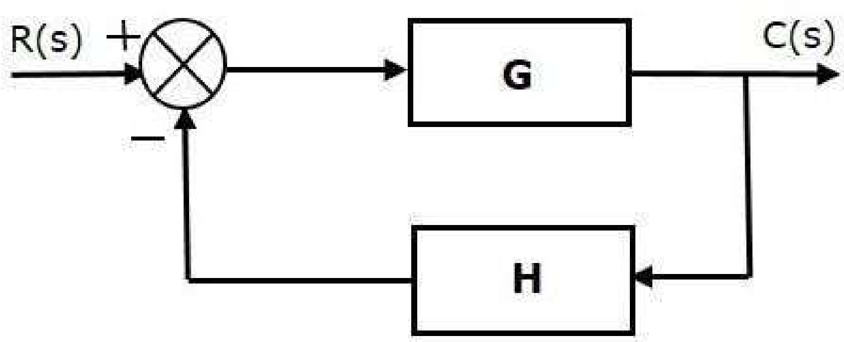
$$T = \frac{G}{1-GH} \qquad \text{(Equation 1)}$$

Where,

- **T** is the transfer function or overall gain of positive feedback control system.
- **G** is the open loop gain, which is function of frequency.
- **H** is the gain of feedback path, which is function of frequency.

Negative Feedback

Negative feedback reduces the error between the reference input, R(s)R(s) and system output. The following figure shows the block diagram of the **negative feedback control system**.



Transfer function of negative feedback control system is,

$$T = \frac{G}{1+GH} \qquad \text{(Equation 2)}$$

Where,

- **T** is the transfer function or overall gain of negative feedback control system.
- **G** is the open loop gain, which is function of frequency.
- **H** is the gain of feedback path, which is function of frequency.

The derivation of the above transfer function is present in later chapters.

Effects of Feedback

Let us now understand the effects of feedback.

Effect of Feedback on Overall Gain

- From Equation 2, we can say that the overall gain of negative feedback closed loop control system is the ratio of 'G' and (1+GH). So, the overall gain may increase or decrease depending on the value of (1+GH).
- If the value of (1+GH) is less than 1, then the overall gain increases. In this case, 'GH' value is negative because the gain of the feedback path is negative.
- If the value of (1+GH) is greater than 1, then the overall gain decreases. In this case, 'GH' value is positive because the gain of the feedback path is positive.

In general, 'G' and 'H' are functions of frequency. So, the feedback will increase the overall gain of the system in one frequency range and decrease in the other frequency range.

Effect of Feedback on Sensitivity

Sensitivity of the overall gain of negative feedback closed loop control system (**T**) to the variation in open loop gain (**G**) is defined as

$$S_G^T = \frac{\frac{\partial T}{T}}{\frac{\partial G}{G}} = \frac{Percentage\ change\ in\ T}{Percentage\ change\ in\ G} \qquad \text{(Equation 3)}$$

Where, ∂T is the incremental change in T due to incremental change in G.

We can rewrite Equation 3 as

$$S_G^T = \frac{\partial T}{\partial G}\frac{G}{T} \qquad \text{(Equation 4)}$$

Do partial differentiation with respect to G on both sides of Equation 2.

$$\frac{\partial T}{\partial G} = \frac{\partial}{\partial G}\left(\frac{G}{1+GH}\right) = \frac{(1+GH).1-G(H)}{(1+GH)^2} = \frac{1}{(1+GH)^2} \qquad \text{(Equation 5)}$$

From Equation 2, you will get

$$\frac{G}{T} = 1 + GH \qquad \text{(Equation 6)}$$

Substitute Equation 5 and Equation 6 in Equation 4.

$$S_G^T = \frac{1}{(1+GH)^2}(1+GH) = \frac{1}{1+GH}$$

So, we got the **sensitivity** of the overall gain of closed loop control system as the reciprocal of (1+GH). So, Sensitivity may increase or decrease depending on the value of (1+GH).

- If the value of (1+GH) is less than 1, then sensitivity increases. In this case, 'GH' value is negative because the gain of feedback path is negative.
- If the value of (1+GH) is greater than 1, then sensitivity decreases. In this case, 'GH' value is positive because the gain of feedback path is positive.

In general, 'G' and 'H' are functions of frequency. So, feedback will increase the sensitivity of the system gain in one frequency range and decrease in the other frequency range. Therefore, we have to choose the values of 'GH' in such a way that the system is insensitive or less sensitive to parameter variations.

Effect of Feedback on Stability

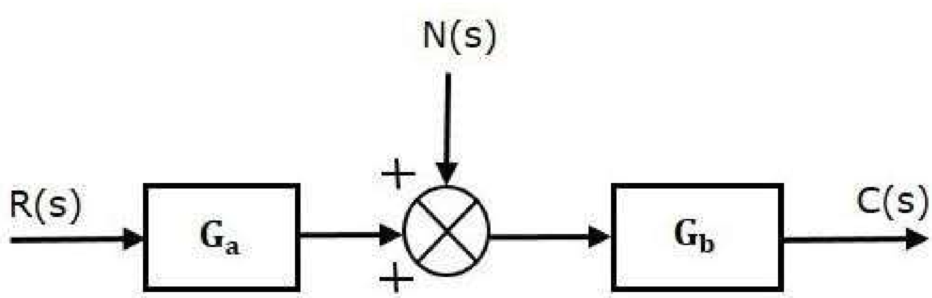
- A system is said to be stable, if its output is under control. Otherwise, it is said to be unstable.
- In Equation 2, if the denominator value is zero (i.e., GH = -1), then the output of the control system will be infinite. So, the control system becomes unstable.

Therefore, we have to properly choose the feedback in order to make the control system stable.

Effect of Feedback on Noise

To know the effect of feedback on noise, let us compare the transfer function relations with and without feedback due to noise signal alone.

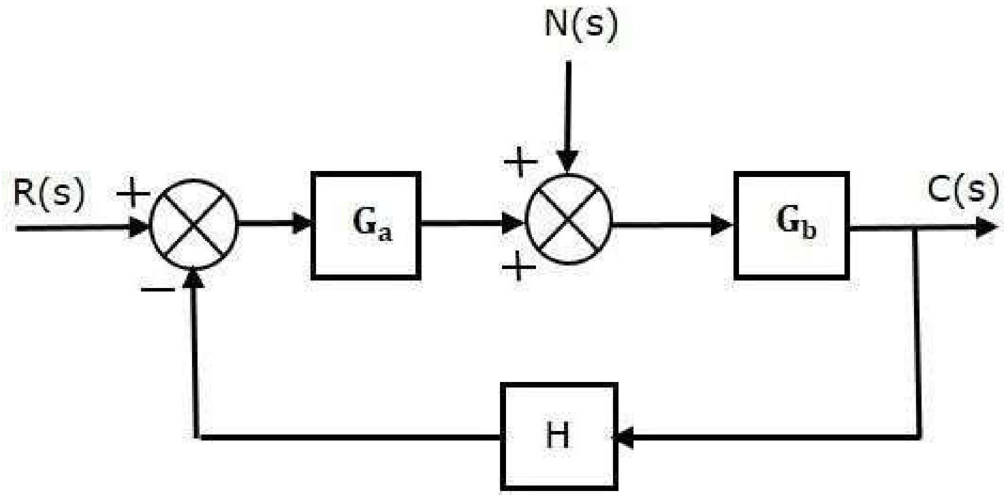
Consider an **open loop control system** with noise signal as shown below.



The **open loop transfer function** due to noise signal alone is

$$\frac{C(s)}{N(s)} = G_b \qquad \text{(Equation 7)}$$

It is obtained by making the other input $R(s)$ equal to zero.



The **closed loop transfer function** due to noise signal alone is

$$\frac{C(s)}{N(s)} = \frac{G_b}{1+G_a G_b H} \qquad \text{(Equation 8)}$$

It is obtained by making the other input $R(s)$ equal to zero.

Compare Equation 7 and Equation 8,

In the closed loop control system, the gain due to noise signal is decreased by a factor of $(1+G_aG_bH)$ provided that the term $(1+G_aG_bH)$ is greater than one.

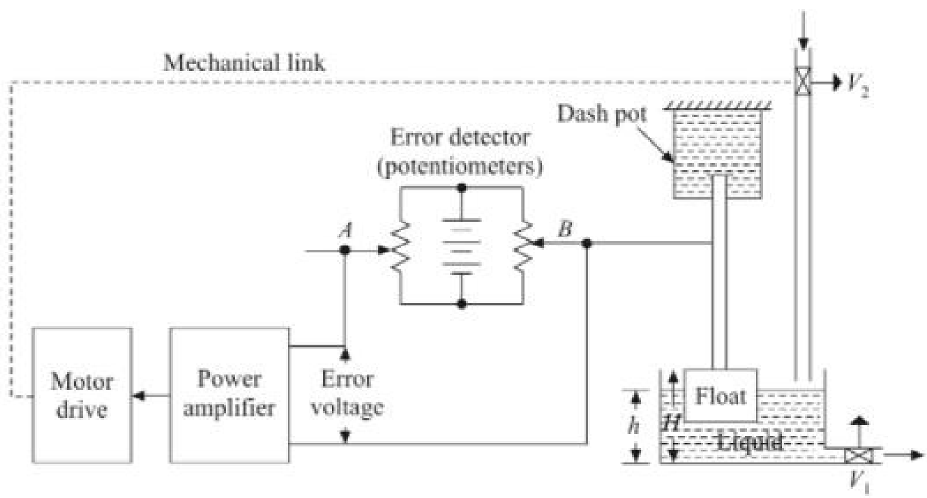
SERVOMECHANISM

In modern usage, the term servomechanism or servo is restricted to feedback controlsystems in which the controlled variable is mechanical position or time derivatives of position, e.g. velocity and acceleration. Few servo mechanisms are illustrated below.

Automatic Tank Level Control System

Figure shows an automatic tank level control system. The purpose of this system is to maintain the liquid level h (output) in the tank as close to the desired liquid level H as possible, even when the output flow rate is varied by opening the valve V_1. This has to be done by controlling the opening of the valve V_2. The potentiometer acts as an error detector. The slider arm A is positioned corresponding to the desired liquid level H (the reference input). The power amplifier and the motor drive form the control elements. The float forms the feedback path element. The valve V_2 to be controlled is the plant.

The liquid level is sensed by a float and it positions the slider arm B on the potentiometer. When the liquid level rises or falls, the potentiometer gives an error voltage proportional to the change in liquid level. The error voltage actuates the motor through a power amplifier which in turn conditions the plant (i.e. decreases or increases the opening of the valve V_2) in order to restore the desired liquid level. Thus, the control system automatically attempts to correct any deviation between the actual and desired liquid levels in the tank.



Automatic tank level control system.

CHAPTER-02

MATHEMATICAL MODEL OF A SYSTEM

The control systems can be represented with a set of mathematical equations knownas **mathematical model**. These models are useful for analysis and design of control systems.

Impulse Response and Transfer Functions

The classical way of modelling linear systems is to use transfer functions to represent input-output relations between variables. One way to define the transfer function is to use the impulse response which is defined as follows:

Impulse response: Consider that a linear time-invariant system has the input $r(t)$ and the output $c(t)$. The system can be characterized by its impulse response $g(t)$, which is defined as the output when the input is a unit impulse function $\delta(t)$. Once the impulse response of a linear system is known, the output of the system, $c(t)$, with any input $r(t)$ can be found by using the transfer function.

Transfer Function

The transfer function $G(s)$ is related to the Laplace transform of the input and the output through the following relation

$$G(s) = \frac{C(s)}{R(s)}$$

with all the initial conditions set to zero, and $C(s)$ and $R(s)$ are the Laplace transforms of $c(t)$ and $r(t)$ respectively. That is the transfer function is defined as the ratio of the Laplace transform of the output to the Laplace transform of the input with all initial conditions neglected.

The properties of the transfer function are as follows:

1. The transfer function is defined only for a linear time-invariant system. It is not defined for nonlinear systems.
2. The transfer function between an input variable and an output variable of a system is defined as the Laplace transform of the impulse response. Alternatively, the transfer function between a pair of input and output variables of a system is the ratio of the Laplace transform of the output to the Laplace transform of the input.
3. All initial conditions of the system are set to zero.
4. The transfer function is independent of the input of the system.
5. The transfer function of a continuous-data system is expressed only as a function of the complex variable s. It is not a function of the real variable time, or any other variable that is used as the independent variable. For discrete-data systems modelled by difference equations, the transfer function is a function of z when the Z transform is used

Advantages of Transfer Function

1. Once transfer function is known, output response for any type of reference input can be calculated.
2. It helps in determining the important information about the system i.e. poles', zeros, characteristic equation etc..
3. It helps in the stability analysis of the system.
4. The system differential equation can be easily obtained by replacing variable 's' by d/dt.
5. Finding inverse, the required variable can be easily expressed in the time domain. This is much more easy than to analyse the entire system in the time domain.

Disadvantages of Transfer Function

The few limitations of the transfer function approach called approach are,

i) Only applicable to linear time invariant systems.

ii) It does not provide any information concerning the physical structure of the system. From transfer function, physical nature of the system whether it is electrical, mechanical, thermal or hydraulic, cannot be judged.

iii) Effects arising due to initial conditions are totally neglected. Hence initial conditions loose their importance.

Poles & Zeroes of transfer Function

$$\text{T.F.} = \frac{P(s)}{Q(s)}$$

This can be further expressed as,

$$= \frac{a_0 s^m + a_1 s^{m-1} + a_2 s^{m-2} + \ldots + a_m}{b_0 s^n + b_1 s^{n-1} + b_2 s^{n-2} + \ldots + b_n}$$

The numerator and denominator can be factorised to get the factorised form of the transfer function as,

$$\text{T.F.} = \frac{K(s - s_a)(s - s_b) \ldots\ldots (s - s_m)}{(s - s_1)(s - s_2) \ldots\ldots (s - s_n)}$$

where K is called **system gain factor.**

values $s_1, s_2, s_3 \ldots\ldots s_n$ are called poles of the T.F.

These poles are nothing but the roots of the equation obtained by equating denominator of a T.F. to zero.

For example, let the transfer function of a system be,

$$T(s) = \frac{2(s+2)}{s(s+4)}$$

The equation obtained by equating denominator to zero is, $s(s + 4) = 0$

$\therefore \quad s = 0 \quad$ and $\quad s = -4$

Similar to the poles, now if the values of 's' are substituted as s_a, s_b ......+ s_m in the numerator of a T.F., its value becomes zero.

Example *The transfer function of a system is given by,*

$$T(s) = \frac{K(s+6)}{s(s+2)(s+5)(s^2+7s+12)}$$

Determine i) Poles ii) Zeros iii) Characteristic equation and iv) Pole-zero plot in s-plane.

SOLUTION:

i) Poles are the roots of the equation obtained by equating denominator to zero i.e. roots of,

$$s(s+2)(s+5)(s^2+7s+12) = 0$$

i.e. $s(s + 2)(s + 5)(s + 3)(s + 4) = 0$

So there are 5 poles located at $s = 0, -2, -5, -3$ and -4

ii) Zeros are the roots of the equation obtained by equating numerator to zero i.e. roots of $K(s + 6) = 0$

i.e. $s = -6$

There is only one zero.

iii) Characteristic equation is one, whose roots are the poles of the transfer function. So it is,

$$s(s+2)(s+5)(s^2+7s+12) = 0$$

i.e. $\quad s(s^2+7s+10)(s^2+7s+12) = 0$

i.e $s^5 + 14s^4 + 71s^3 + 154s^2 + 120s = 0$

iv) Pole-zero plot

This is shown in the Fig.

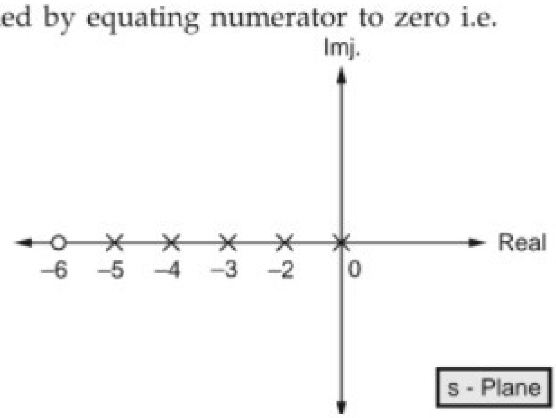


Fig.

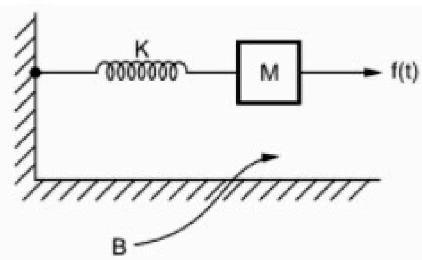
ANALOGOUS SYSTEMS

Comparing equations for the mechanical translational system or for the mechanical rotational system and for the series electrical system, it is seen that they are of identical form. Such systems whose differential equations are of identical form are called *analogous systems*. The force F (torque T) and voltage e are the analogous variables here. This is called the force (torque)-voltage analogy. A list of analogous variables in this analogy is given in Table

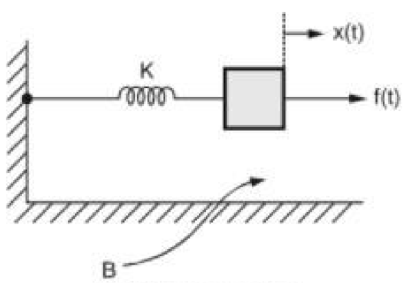
Table Analogous quantities in force (torque)-voltage analogy

Mechanical translational system	*Mechanical rotational system*	*Electrical system*
Force F	Torque T	Voltage e
Mass M	Moment of inertia J	Inductance L
Viscous friction coefficient f	Viscous friction coefficient f	Resistance R
Spring stiffness K	Torsional spring stiffness K	Reciprocal of capacitance $1/C$
Displacement x	Angular displacement θ	Charge q
Velocity v	Angular velocity ω	Current i

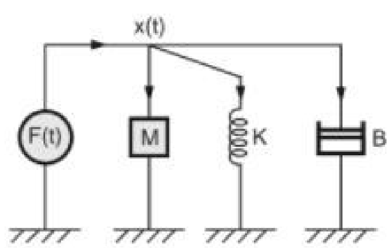
Example *For the physical system shown draw its equivalent system and write equilibrium equations. Hence draw its electrical analogous circuits based on*
i) Force -Voltage ii) Force - Current method.



Solution : Mass `M' will displace by amount `x' and as spring is connected to fixed support and friction `B' is also with respect to fixed support, both K and B will be under influence of `x' only. Now its equivalent system will contain one node and as all elements are under influence of x(t) alone, must be connected in parallel under that node.



(a) Physical system



(b) Equivalent mechanical system

The equilibrium equation will be,

$$f(t) = M\frac{d^2 x(t)}{dt^2} + K\,x(t) + B\frac{dx(t)}{dt}$$

Taking Laplace, $F(s) = Ms^2 X(s) + K X(s) + Bs X(s) = X(s)\left[Ms^2 + K + Bs\right]$ ... (1)

i) Force-Voltage Method : Use the analogous terms as,

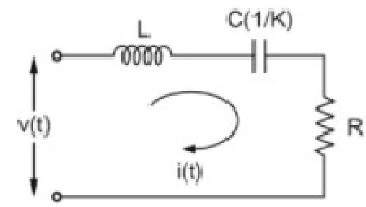
$$M \to L,\ B \to R,\ K \to \frac{1}{C},\ x \to q,\ \frac{dx}{dt} \to \frac{dq}{dt} \to i,\ x \to \int idt,\ \frac{d^2x}{dt^2} = \frac{di}{dt}$$

All quantities are expressed in terms of current i.

$$v(t) = L\frac{di}{dt} + \frac{1}{C}\int idt + Ri \quad \text{... (2)}$$

$$\therefore \quad V(s) = sL\,I(s) + \frac{I(s)}{sC} + I(s)R \quad \text{... (3)}$$

Simulate using loop method :



Analogous to K is a capacitor C but its value is proportional to 1/K hence it is indicated by writing (1/K) in the bracket near C. This is shown in the Fig.

ii) Force-Current Method : Use the analogous terms as,

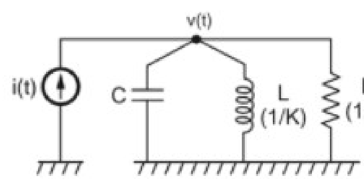
$$M \to C,\ B \to \frac{1}{R},\ K \to \frac{1}{L},\ x \to \phi,\ \frac{dx}{dt} \to \frac{d\phi}{dt} \to v(t),\ x \to \int v(t)dt,\ \frac{d^2x}{dt^2} = \frac{dv(t)}{dt}$$

All quantities are expressed in terms of voltage v.

$$\therefore \quad i(t) = C\frac{dv(t)}{dt} + \frac{1}{R}v(t) + \frac{1}{L}\int v(t)dt \quad \text{... (4)}$$

$$\therefore \quad I(s) = sCV(s) + \frac{1}{R}V(s) + \frac{1}{sL}V(s) \quad \text{... (5)}$$

Simulate using node method :



Analogous to K is an inductor L while to B is a resistor R. But their values are proportional to reciprocals of K and B respectively. This is indicated by writing (1/K) and (1/B) in the brackets near L and R respectively in the Fig.

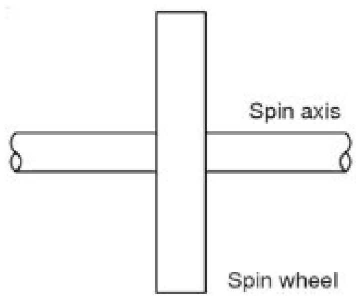
Gyroscope

Gyroscope is used in the navigation system and as gyroscope works on the inertia principle it is called inertial navigation system.

CONSTRUCTION OF THE GYROSCOPE

The Gyroscope works on the inertial principle. Inertia of the mechanical system is similar to inductance of the electrical system. In electrical system, the change of current or voltage is opposed by the inductance of the system. In mechanical system the change of position is opposed by the inertia. According to Newton's law of inertia, if any object is stationary, it will remain stationary unless acted upon by a force. If it is in motion, it will remain in motion. In linear motion, force is required to overcome the inertia given by the mass. In case of angular motion, torque is required to impart change of angular motion overcoming rotational moment of inertia.

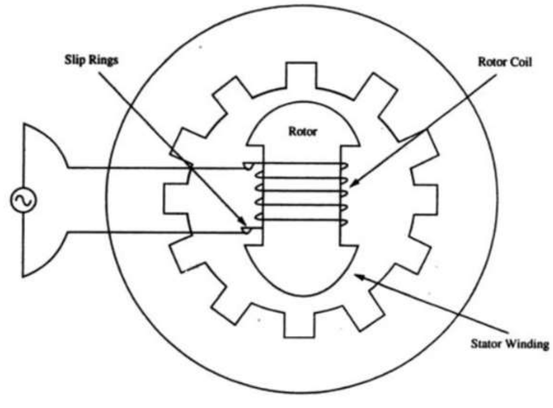
In case of gyroscope, this inertia comes from the kinetic energy stored and momentum attained by the rotating element. For this purpose the gyroscope has a spin wheel. It has a small thickness and large diameter, because the rotational moment of inertia of the disc is proportional to the forth power of the radius. To allow the rotation of a spin wheel it has an axis of rotation provided by a shaft on which the spin wheel is mounted as shown in Figure



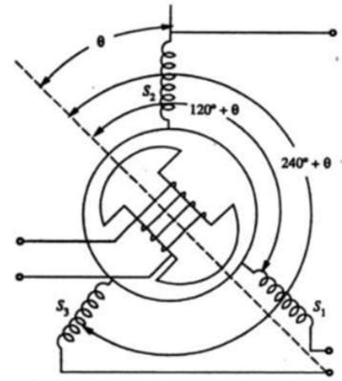
Synchros

A synchro (selsyn or autosyn) is an electromagnetic transducer that is used to convert an angular shaft position into an electric signal. The basic element of the synchro is a synchro transmitter whose construction is very similar to that of a three-phase alternator. The stator which is stationary is of laminated silicon steel and is slotted to accommodate a balanced three-phase winding which is of concentric coil type in which three identical coils are placed in the stator with their axis 120° apart and is Y-connected.The rotor is of dumbbell construction and is wound with a concentric coil. Through the slip rings, an a.c. voltage is applied to the rotor winding.

The constructional features and schematic diagram of a synchro transmitter are shown in Fig.



Constructional features of synchro transmitter.



Schematic diagram of synchro transmitter.

Let v_{s_1}, v_{s_2} and v_{s_3} be the voltages induced in the stator coils S_1, S_2 and S_3 with respect to the neutral respectively. Then, for the rotor position of the synchro transmitter shown in Fig. where the rotor axis makes an angle θ with the axis of the stator coil S_2

$$v_{s_1} = KV_r \sin \omega_c t \ . \ \cos (\theta + 120°)$$

$$v_{s_2} = KV_r \sin \omega_c t \ . \ \cos \theta$$

$$v_{s_3} = KV_r \sin \omega_c t \ . \ \cos (\theta + 240°)$$

The three terminal voltages of the stator are

$$v_{s_1 s_2} = v_{s_1} - v_{s_2} = \sqrt{3}\, K\, V_r \sin (\theta + 240°) \sin \omega_c t$$

$$v_{s_2 s_3} = v_{s_2} - v_{s_3} = \sqrt{3}\, KV_r \sin (\theta + 120°) \sin \omega_c t$$

$$v_{s_3 s_1} = v_{s_3} - v_{s_1} = \sqrt{3}\, KV_r \sin \theta \ . \ \sin \omega_c t$$

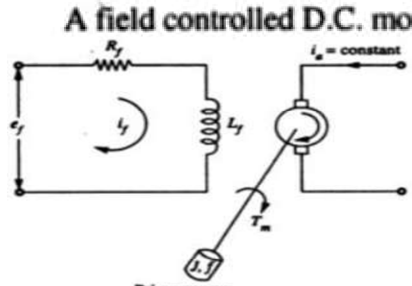
DC servomotors

While deriving transfer functions, the system is approximated by a linear lumped constant parameters model by making suitable assumptions.

As an example to illustrate this, let us derive the transfer function of a d.c. servomotor. In servo applications, a d.c. motor is expected to produce very high acceleration from standstill. The requirements of the d.c. motor are low inertia and high starting torque. A low value of inertia can be attained by reducing the armature diameter in order that the desired output power is achieved. A d.c. servomotor is an ordinary d.c. motor but for some minor changes. The two different modes in which the d.c. motor can be operated are (*i*) field control mode and (*ii*) armature control mode.

(*a*) Transfer Function of a Field Controlled D.C. Motor

A field controlled D.C. motor is shown in Fig.



Field Controlled D.C. Motor.

Let R_f and L_f be the resistance and inductance of the field circuit with excitation voltage e_f and current i_f. Let T_m be the torque developed by the motor in N-m. Let 'J' be the moment of inertia in kg-m^2, 'f' be the coefficient of friction in N-m/(rad/sec) and 'θ' be the angular displacement of motor shaft in radians.

Applying Kirchhoff's law for the field circuit

$$e_f(t) = R_f i_f(t) + L_f \frac{di_f(t)}{dt}$$

Taking Laplace Transform of equation (5.19), we have,

$$E_f(s) = (R_f + sL_f)\; I_f(s)$$

$$I_f(s) = \frac{E_f(s)}{(R_f + sL_f)}$$

The torque developed is proportional to the field current, since the armature current is constant.

$$T_m(t) = K_t i_f(t)$$

where K_t is a constant.

Taking Laplace Transform of equation (5.21),

$$T_m(s) = K_t \,.\, I_f(s)$$

The torque equation is given by

$$T_m(t) = J\frac{d^2\theta}{dt^2} + f\frac{d\theta}{dt}$$

Taking Laplace Transform,

$$T_m(s) = (Js^2 + fs)\;\theta(s)$$

$$(Js^2 + fs)\,\theta(s) = K_t \,.\, I_f(s)$$

$$s\,(Js + f)\;\theta(s) = K_t \cdot \frac{E_f(s)}{(R_f + sL_f)}$$

The transfer function is given by

$$G(s) = \frac{\theta(s)}{E_f(s)}$$

$$= \frac{K_t}{s\,(Js + f)\;(R_f + sL_f)}$$

$$= \frac{K_t}{fs\left[1 + s\left(\frac{J}{f}\right)\right] R_f\left[1 + s\left(\frac{L_f}{R_f}\right)\right]}$$

$$G(s) = \frac{K_t'}{s\,(1 + \tau_{me})\;(1 + s\,\tau_f)}$$

where

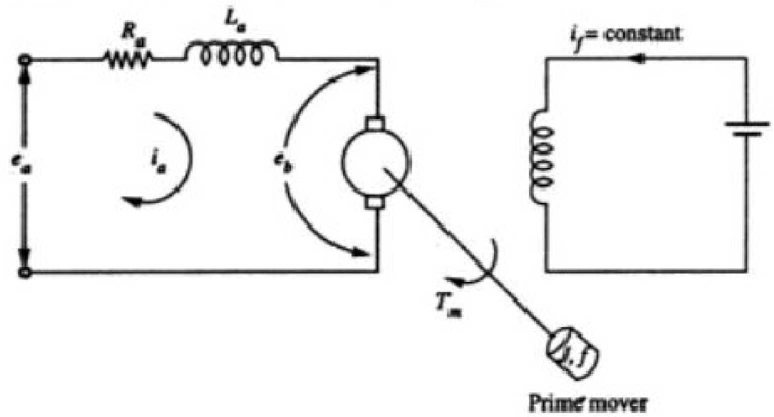
$$K_t' = \frac{K_t}{R_f - f}$$

$$\tau_{me} = \frac{J}{f} = \text{Mechanical time constant}$$

$$\tau_f = \frac{L_f}{R_f} = \text{field time constant}$$

(b) Transfer Function of an Armature Controlled D.C. Motor

A D.C. armature controlled motor is shown in Fig.



Armature Controlled D.C. Motor.

Let R_a and L_a be the resistance and inductance of the armature circuit with an applied voltage of e_a. Let e_b be the back e.m.f. in the armature circuit. Let i_f be the constant field current through the field windings. Let T_m be the torque developed in N-m, 'J' be the moment of inertia in kg-m^2, 'f' be the friction coefficient in $\frac{\text{N–m}}{\text{rad/sec}}$ and θ, the angular displacement in radians.

Applying Kirchhoff's voltage law to the armature circuit,

$$e_a(t) = R_a\, i_a(t) + L_a \cdot \frac{di_a(t)}{dt} + e_b(t)$$

Taking Laplace Transform of equation , we have,

$$E_a(s) = [R_a + sL_a]\ I_a(s) + E_b(s)$$

$$I_a(s) = \frac{E_a(s) - E_b(s)}{(R_a + sL_a)}$$

Since the field current is kept constant, the torque developed is proportional to the armature current, *i.e.*,

$$T_m(t) = K_t\, i_a(t)$$

Taking Laplace Transform of equation (5.28), we have,

$$T_m(s) = K_t\, I_a(s)$$

The torque equation is given by

$$T_m(t) = J\frac{d^2\theta(t)}{dt^2} + f\frac{d\theta(t)}{dt}$$

Taking Laplace Transform of equation

$$T_m(s) = Js^2\theta(s) + fs\theta(s)$$

$$\theta(s)(sf + s^2 J) = K_t I_a(s)$$

$$s\theta(s)[f+sJ] = K_t\frac{[E_a(s) - E_b(s)]}{(R_a + sL_a)}$$

Motor back e.m.f. is proportional to speed, *i.e.*,

$$e_b(t) = K_b \cdot \frac{d\theta}{dt}$$

where K_b = back emf = constant.

Hence $$E_b(s) = sK_b \cdot \theta(s)$$

Therefore, $$s\theta(s)(f+sJ) = \frac{K_t[E_a(s) - sK_b \cdot \theta(s)]}{(R_a + sL_a)}$$

$$s\theta(s)(f+sJ)(R_a+sL_a)+sK_t \cdot K_b \cdot \theta(s) = K_t \cdot E_a(s)$$

Transfer function is given by

$$G(s) = \frac{\theta(s)}{E_a(s)}$$

$$= \frac{K_t}{sK_t K_b + s(f+sJ)(R_a+sL_a)}$$

$$G(s) = \frac{K_t}{s[K_t K_b + f(1 + s\tau_{me})R_a(1 + s\tau_a)]}$$

where $$\tau_{me} = \text{Mechanical time constant} = \frac{J}{f}$$

$$\tau_a = \text{armature time constant} = \frac{L_a}{R_a}$$

Hence $$G(s) = \frac{K_t}{s[K_t K_b + R_a f(1+s\tau_{me})(1+s\tau_a)]}$$

Usually for an armature controlled machine, $\tau_a = 0$.

Hence we can have,

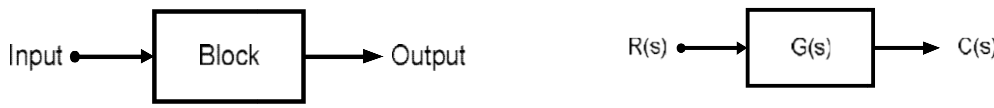
$$G(s) = \frac{K_t}{s[K_t K_b + fR_a(1+s\tau_{me})]}$$

CHAPTER-04

BLOCK DIAGRAM ALGEBRA & SIGNAL FLOW GRAPHS

BASIC DEFINITION IN BLOCK DIAGRAM MODEL:

Block diagram: It is the pictorial representation of the cause-and-response relationship between input and output of a physical system.



(a) (b)

Fig. (a) A block diagram representation of a system and
(b) A block diagram representation with gain of a system

Output: The value of input multiplied by the gain of the system.

$$C(s)=G(s)R(s)$$

Summing point: It is the component of a block diagram model at which two or more signals can be added or subtracted. In Fig, inputs R(s) and B(s) have been given to a summing point and its output signal is E(s). Here,

$$E(s)=R(s)-B(s)$$

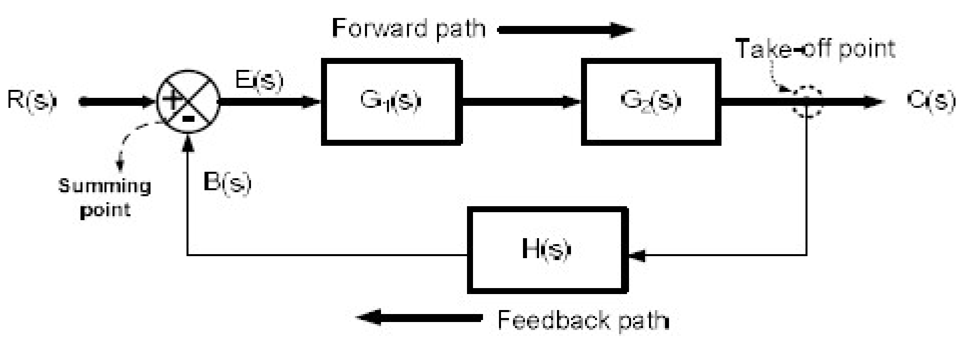


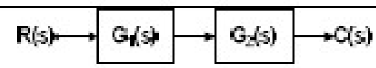
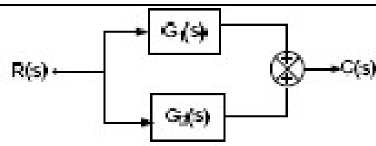
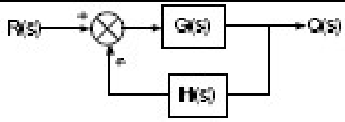
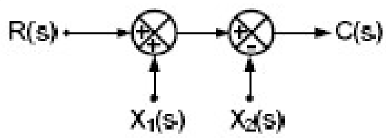
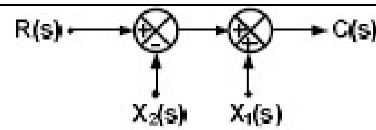
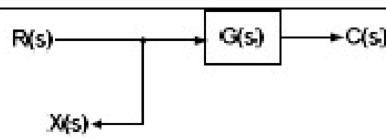
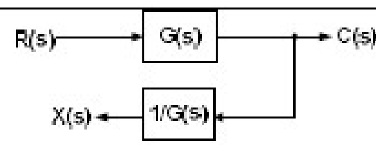
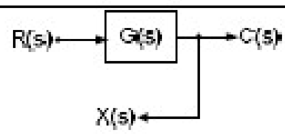
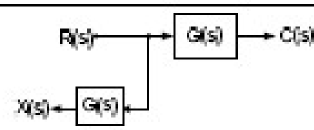
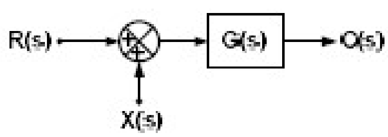
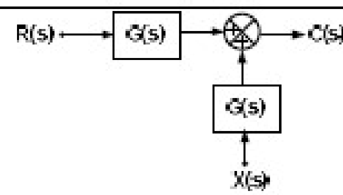
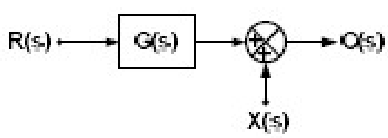
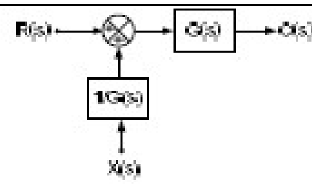
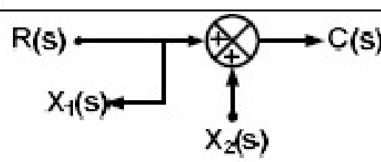
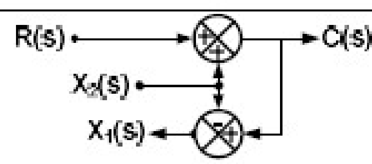
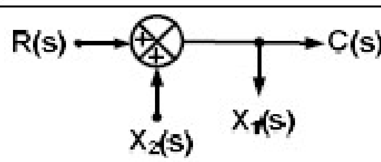
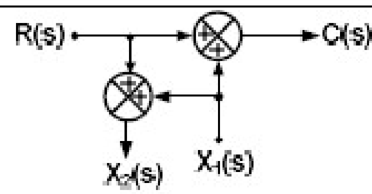
Fig. A block diagram representation of a system showing its different components

Take-off point: It is the component of a block diagram model at which a signal can be taken directly and supplied to one or more points as shown in Fig.

Forward path: It is the direction of signal flow from input towards output.

Feedback path: It is the direction of signal flow from output towards input.

RULES FOR REDUCTION OF BLOCK DIAGRAM MODEL:

Sl. No.	Rule No.	Configuration	Equivalent	Name
1	Rule 1			Cascade
2	Rule 2			Parallel
3	Rule 3			Loop
4	Rule 4			Associative Law
5	Rule 5			Move take-off point after a block
6	Rule 6			Move take-off point before a block
7	Rule 7			Move summing-point point after a block
8	Rule 8			Move summing-point point before a block
9	Rule 9			Move take-off point after a summing-point
10	Rule 10			Move take-off point before a summing-point

PROCEDURE FOR REDUCTION OF BLOCK DIAGRAM MODEL:

Step 1: Reduce the cascade blocks.

Step 2: Reduce the parallel blocks.

Step 3: Reduce the internal feedback loops.

Step 4: Shift take-off points towards right and summing points towards left.

Step 5: Repeat step 1 to step 4 until the simple form is obtained.

Step 6: Find transfer function of whole system as $\frac{C(s)}{R(s)}$.

PROCEDURE FOR FINDING OUTPUT OF BLOCK DIAGRAM MODEL WITH MULTIPLE INPUTS:

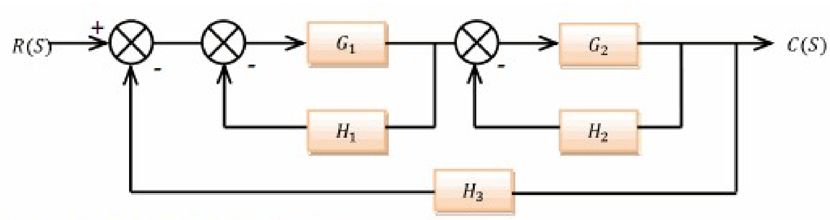
Step 1: Consider one input taking rest of the inputs zero, find output.

Step 2: Follow step 1 for each inputs of the given Block Diagram model and find their corresponding outputs.

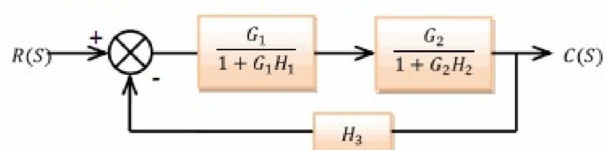
Step 3: Find the resultant output by adding all individual outputs.

Example 1:- Find C(s)/R(s) using block diagram reduction rules

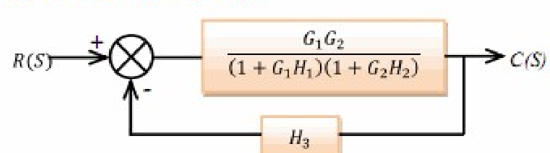
(1)



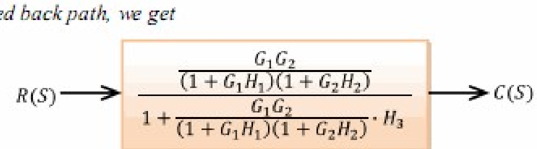
olution: By eliminating the feed-back paths, we get

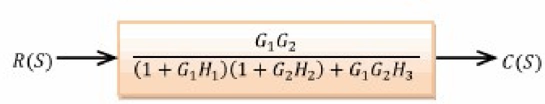


Combining the blocks in series, we get



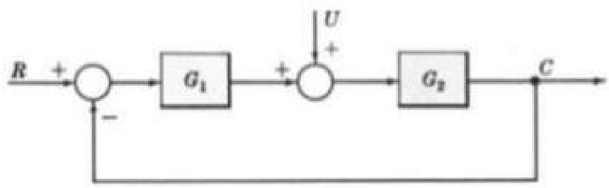
Eliminating the feed back path, we get



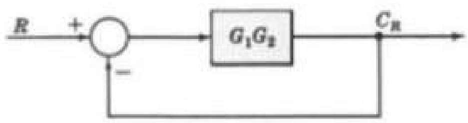


$$\Rightarrow \boxed{TF = \frac{C(S)}{R(S)} = \frac{G_1G_2}{(1+G_1H_1)(1+G_2H_2)+G_1G_2H_3}}$$

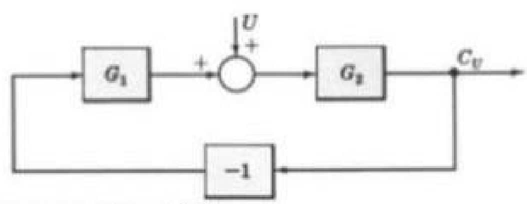
Example-2:- Determine output C due to inputs R & U using the superposition method.



Step 1: Put $U \equiv 0$.
Step 2: The system reduces to

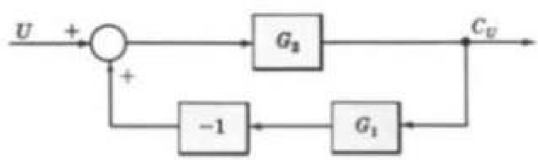


Step 3: the output C_R due to input R is $C_R = [G_1G_2/(1 + G_1G_2)]R$.

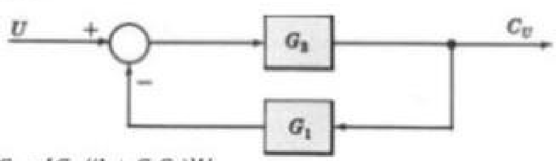


Step 4a: Put $R = 0$.
Step 4b: Put -1 into a block, representing the negative feedback effect:

Rearrange the block diagram:



Let the -1 block be absorbed into the summing point:



Step 4c: the output C_U due to input U is $C_U = [G_2/(1 + G_1G_2)]U$.

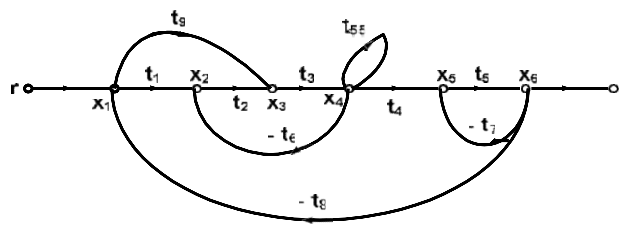
Step 5: The total output is $C = C_R + C_U$

$$= \left[\frac{G_1G_2}{1 + G_1G_2}\right]R + \left[\frac{G_2}{1 + G_1G_2}\right]U$$

$$= \left[\frac{G_2}{1 + G_1G_2}\right][G_1R + U]$$

SIGNAL FLOW GRAPHS (SFGS)

It is a pictorial representation of a system that graphically displays the signal transmission in it.



Basic Definitions in SFGs:

Input or source node: It is a node that has only outgoing branches i.e. node ‘r’.

Output or sink node: It is a node that has only incoming branches i.e. node ‘c’.

Chain node: It is a node that has both incoming and outgoing branches i.e. nodes ‘x_1’, ‘x_2’,‘x_3’,‘x_4’,‘x_5’and ‘x_6’.

Gain or transmittance: It is the relationship between variables denoted by two nodes or value of a branch., Transmittances are ‘t_1’, ‘t_2’,‘t_3’,‘t_4’,‘t_5’and ‘t_6’.

Forward path: It is a path from input node to output node without repeating any of the nodes in between them. There are two forward paths, i.e. path-1:‘r-x_1-x_2-x_3-x_4-x_5-x_6-c’ and path-2:‘r-x_1-x_3-x_4-x_5-x_6-c’.

Feedback path: It is a path from output node or a node near output node to a node near input node without repeating any of the nodes in between them.

Loop: It is a closed path that starts from one node and reaches the same node after trading through other nodes. There are four loops, i.e. loop-1:‘x_2-x_3-x_4-x_1’, loop-2:‘x_5-x_6- x_5’, loop-3:‘x_1-x_2-x_3-x_4-x_5-x_6-x_1’ and loop-4:‘x_1-x_3-x_4-x_5-x_6-x_1’.

Self Loop: It is a loop that starts from one node and reaches the same node without trading through other nodes i.e. loop in node ‘x_4’ with transmittance ‘t_{55}’.

Path gain: It is the product of gains or transmittances of all branches of a forward path. The path gains are $P_1 = t_1t_2t_3t_4t_5$ (for path-1) and $P_2 = t_9t_3t_4t_5$ (for path-2).

Loop gain: It is the product of gains or transmittances of all branches of a loop. There are four loops, i.e. $L_1 = -t_2t_3t_6$, $L_2 = -t_5t_7$, $L_3 = -t_1t_2t_3t_4t_5t_8$, and $L_4 = -t_9t_3t_4t_5t_8$.

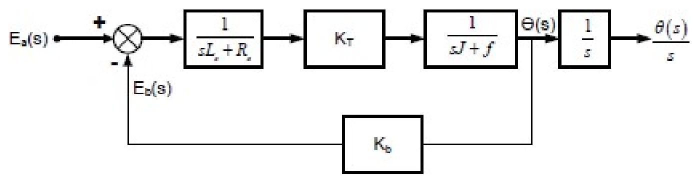
Dummy node: If the first node is not an input node and/or the last node is not an output node than a node is connected before the existing first node and a node is connected after the existing last node with unity transmittances. These nodes are called dummy nodes. ‘r’ and ‘c’ are the dummy nodes.

Non-touching Loops: Two or more loops are non-touching loops if they don’t have any common nodes between them. L_1 and L_2 are non-touching loops

PROPERTIES OF SFGS:

- Applied to linear system
- Arrow indicates signal flow
- Nodes represent variables, summing points and take-off points
- Algebraic sum of all incoming signals and outgoing nodes is zero
- SFG of a system is not unique
- Overall gain of an SFG can be determined by using Mason’s gain formula

SFG FROM BLOCK DIAGRAM MODEL:

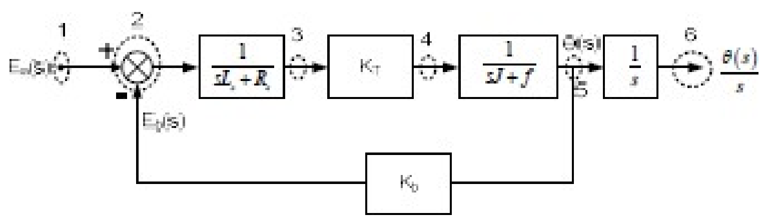


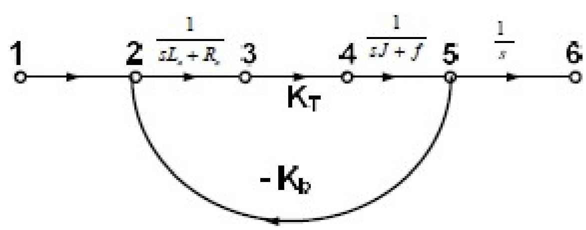
Step-1: All variables and signals are replaced by nodes.

Step-2: Connect all nodes according to their signal flow.

Step-3: Each of gains is replaced by transmittances of the branches connected between two nodes of the forward paths.

Step-4: Each of gains is replaced by transmittances multiplied with (-1) of the branches connected between two nodes of the forward paths.





MASON'S GAIN FORMULA:

Transfer function of a system=

$$G(s) = \frac{C(s)}{R(s)} = \frac{\sum_{k=1}^{N} P_k \Delta_k}{\Delta}$$

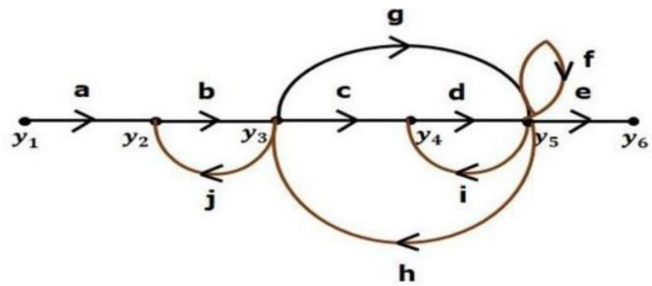
N= total number of forwardpaths

P_k= path gain of k^{th} forward path

Δ= 1 - ($\sum$loop gains of all individual loops) + ($\sum$gain product of loop gains of all possible two non-touching loops) - ($\sum$gain product of loop gains of all possible three non-touching loops) + …

Δ_k= value of Δ after eliminating all loops that touches k^{th} forward path

Example:- Find overall transfer function of system using Mason's gain formula



- Number of forward paths, N = 2.
- First forward path is - $y_1 \rightarrow y_2 \rightarrow y_3 \rightarrow y_4 \rightarrow y_5 \rightarrow y_6$.
- First forward path gain, $p_1 = abcde$.
- Second forward path is - $y_1 \rightarrow y_2 \rightarrow y_3 \rightarrow y_5 \rightarrow y_6$.
- Second forward path gain, $p_2 = abge$.
- Number of individual loops, L = 5.
- Loops are - $y_2 \rightarrow y_3 \rightarrow y_2$, $y_3 \rightarrow y_5 \rightarrow y_3$, $y_3 \rightarrow y_4 \rightarrow y_5 \rightarrow y_3$,

  $y_4 \rightarrow y_5 \rightarrow y_4$ and $y_5 \rightarrow y_5$.
- Loop gains are - $l_1 = bj$, $l_2 = gh$, $l_3 = cdh$, $l_4 = di$ and $l_5 = f$.
- Number of two non-touching loops = 2.
- First non-touching loops pair is - $y_2 \rightarrow y_3 \rightarrow y_2$, $y_4 \rightarrow y_5 \rightarrow y_4$.
- Gain product of first non-touching loops pair, $l_1 l_4 = bjdi$
- Second non-touching loops pair is - $y_2 \rightarrow y_3 \rightarrow y_2$, $y_5 \rightarrow y_5$.
- Gain product of second non-touching loops pair is - $l_1 l_5 = bjf$

Higher number of (more than two) non-touching loops are not present in this signal flow graph.

We know,

$$\Delta = 1 - (sum\ of\ all\ individual\ loop\ gains)$$
$$+(sum\ of\ gain\ products\ of\ all\ possible\ two\ nontouching\ loops)$$
$$-(sum\ of\ gain\ products\ of\ all\ possible\ three\ nontouching\ loops) + \ldots$$

Substitute the values in the above equation,

$$\Delta = 1 - (bj + gh + cdh + di + f) + (bjdi + bjf) - (0)$$

$$\Rightarrow \Delta = 1 - (bj + gh + cdh + di + f) + bjdi + bjf$$

There is no loop which is non-touching to the first forward path.

So, $\Delta_1 = 1$.

Similarly, $\Delta_2 = 1$. Since, no loop which is non-touching to the second forward path.

Substitute, N = 2 in Mason's gain formula

$$T = \frac{C(s)}{R(s)} = \frac{\Sigma_{i=1}^{2} P_i \Delta_i}{\Delta}$$

$$T = \frac{C(s)}{R(s)} = \frac{P_1 \Delta_1 + P_2 \Delta_2}{\Delta}$$

Substitute all the necessary values in the above equation.

$$T = \frac{C(s)}{R(s)} = \frac{(abcde)1 + (abge)1}{1 - (bj + gh + cdh + di + f) + bjdi + bjf}$$

$$\Rightarrow T = \frac{C(s)}{R(s)} = \frac{(abcde) + (abge)}{1 - (bj + gh + cdh + di + f) + bjdi + bjf}$$

Therefore, the transfer function is -

$$T = \frac{C(s)}{R(s)} = \frac{(abcde) + (abge)}{1 - (bj + gh + cdh + di + f) + bjdi + bjf}$$

CONSTRUCTION OF SIGNAL FLOW GRAPH FROM ALGEBRAIC EQUATIONS:-

Let us construct a signal flow graph by considering the following algebraic equations

$$y_2 = a_{12}y_1 + a_{42}y_4$$

$$y_3 = a_{23}y_2 + a_{53}y_5$$

$$y_4 = a_{34}y_3$$

$$y_5 = a_{45}y_4 + a_{35}y_3$$

$$y_6 = a_{56}y_5$$

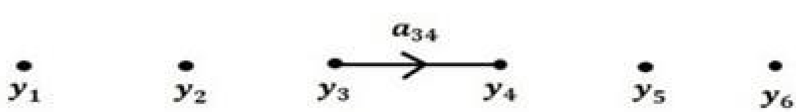
There will be six **nodes** (y_1, y_2, y_3, y_4, y_5 and y_6) and eight **branches** in this signal flow graph. The gains of the branches are a_{12}, a_{23}, a_{34}, a_{45}, a_{56}, a_{42}, a_{53} and a_{35}.

To get the overall signal flow graph, draw the signal flow graph for each equation, then combine all these signal flow graphs and then follow the steps given below −

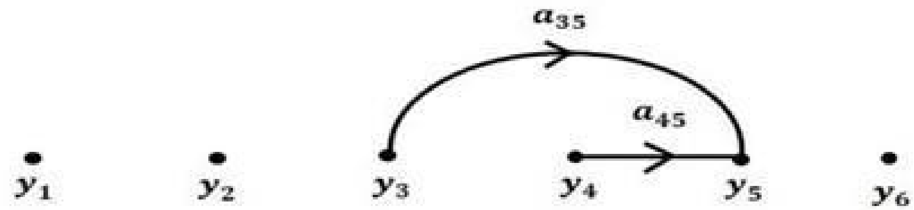
Step 1 − Signal flow graph for $y_2 = a_{13}y_1 + a_{42}y_4$ is shown in the following figure.

Step 2 − Signal flow graph for $y_3 = a_{23}y_2 + a_{53}y_5$ is shown in the following figure.

Step 3 − Signal flow graph for $y_4 = a_{34}y_3$ is shown in the following figure.

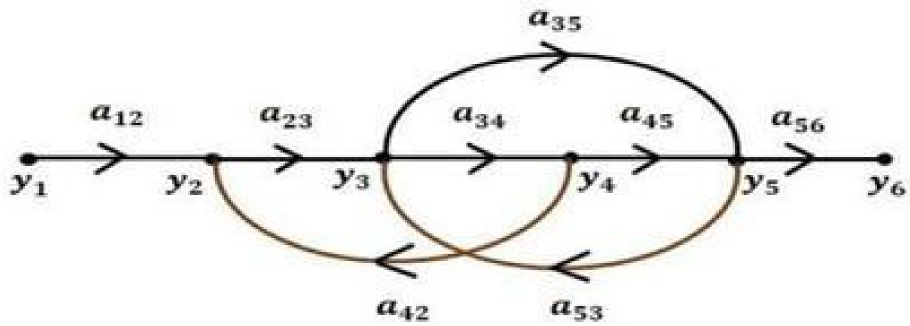


Step 4 – Signal flow graph for $y_5 = a_{45}y_4 + a_{35}y_3$ is shown in the following figure.



Step 5 – Signal flow graph for $y_6 = a_{56}y_5$ is shown in the following figure.

Step 6 – Signal flow graph of overall system is shown in the following figure.

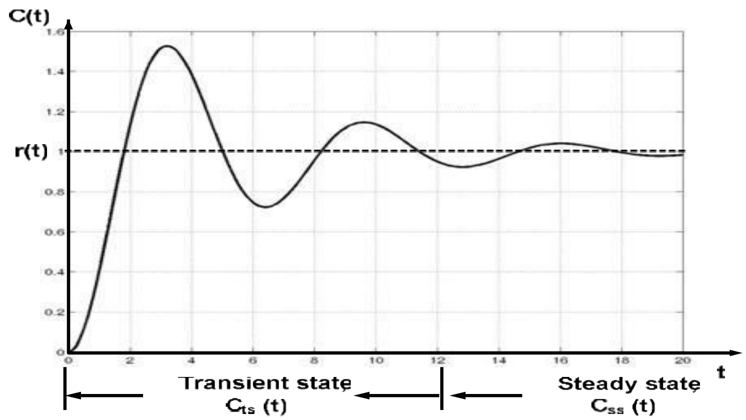


CHAPTER-05

TIME RESPONSE ANALYSIS

TIME RESPONSE OF CONTROL SYSTEM:

Time response c(t)is the variation of output with respect to time. The part of time response that goes to zero after large interval of time is called transient response $c_{tr}(t)$. The part of time response that remains after transient response is called steady-state response $c_{ss}(t)$.



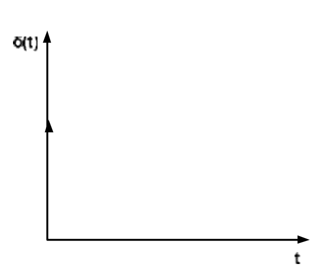
STANDARD TEST SIGNALS

1. Impulse Signal: An impulse signal δ(t) is mathematically defined as follows.

$$\delta(t) = \left.\begin{matrix} undefined & ;t=0 \\ 0 & ;t\neq 0 \end{matrix}\right\}$$

Laplace transform of impulse signal is

$$\delta(s) = 1$$

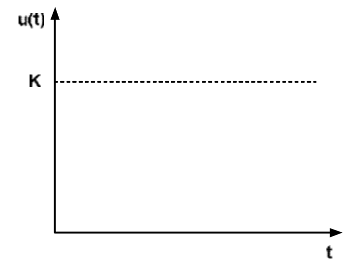


2. Step Signal: A step signal u(t) is mathematically defined as follows.

$$u(t) = \left.\begin{matrix} 0 & ;t<0 \\ K & ;t\geq 0 \end{matrix}\right\}$$

Laplace transform of step signal is

$$U(s) = \frac{K}{s}$$



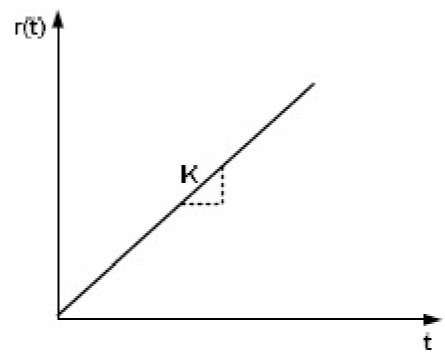
3. Ramp Signal: A step signal r(t) is mathematically defined as follows.

$$r(t) = \left.\begin{matrix} 0 & ;t<0 \\ Kt & ;t \geq 0 \end{matrix}\right\} \qquad (8.10)$$

Laplace transform of ramp signal

is

$$R(s) = \frac{K}{s^2}$$

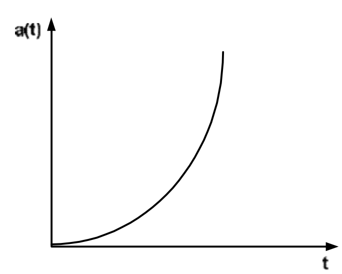


4. Parabolic Signal A step signal a(t) is mathematically defined as follows.

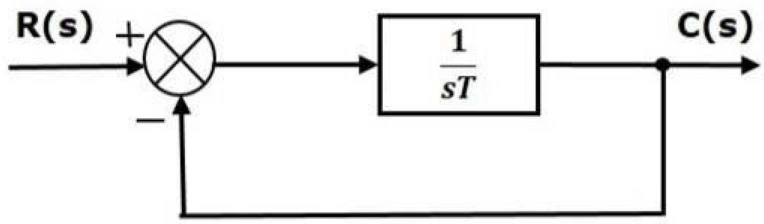
$$a(t) = \left.\begin{matrix} 0 & ;t<0 \\ \frac{Kt^2}{2} & ;t \geq 0 \end{matrix}\right\}$$

Laplace transform of parabolic signal is

$$A(s) = \frac{K}{s^3}$$



TIME RESPONSE OF 1ST ORDER SYSTEM:



We know that the transfer function of the closed loop control system has unity negative feedback as,

$$\frac{C(s)}{R(s)} = \frac{G(s)}{1+G(s)}$$

Substitute, $G(s) = \frac{1}{sT}$ in the above equation.

$$\frac{C(s)}{R(s)} = \frac{\frac{1}{sT}}{1+\frac{1}{sT}} = \frac{1}{sT+1}$$

The power of s is one in the denominator term. Hence, the above transfer function is of the first order and the system is said to be the **first order system**.

We can re-write the above equation as

$$C(s) = \left(\frac{1}{sT+1}\right)R(s)$$

Where,

- **C(s)** is the Laplace transform of the output signal c(t),
- **R(s)** is the Laplace transform of the input signal r(t), and
- **T** is the time constant.

Follow these steps to get the response (output) of the first order system in the time domain.

- Take the Laplace transform of the input signal $r(t)$.
- Consider the equation, $C(s) = \left(\frac{1}{sT+1}\right)R(s)$

- Substitute $R(s)$ value in the above equation.
- Do partial fractions of $C(s)$ if required.
- Apply inverse Laplace transform to $C(s)$.

(i) Unit Step Response:-

Consider the **unit step signal** as an input to first order system.

So, r(t)=u(t)

$$R(s) = \frac{1}{s}$$

Consider the equation, $C(s) = \left(\frac{1}{sT+1}\right) R(s)$

Substitute, $R(s) = \frac{1}{s}$ in the above equation.

$$C(s) = \left(\frac{1}{sT+1}\right)\left(\frac{1}{s}\right) = \frac{1}{s\,(sT+1)}$$

Do partial fractions of C(s).

$$C(s) = \frac{1}{s\,(sT+1)} = \frac{A}{s} + \frac{B}{sT+1}$$

$$\Rightarrow \frac{1}{s\,(sT+1)} = \frac{A\,(sT+1)+Bs}{s\,(sT+1)}$$

On both the sides, the denominator term is the same. So, they will get cancelled by each other. Hence, equate the numerator terms.

$$1=A(sT+1)+Bs$$

By equating the constant terms on both the sides, you will get A = 1.
Substitute, A = 1 and equate the coefficient of the **s** terms on both the sides.

$$0=T+B$$
$$\Rightarrow B=-T$$

Substitute, A = 1 and B = −T in partial fraction expansion of C(s)

$$C(s) = \frac{1}{s} - \frac{T}{sT+1} = \frac{1}{s} - \frac{T}{T\left(s+\frac{1}{T}\right)}$$

$$\Rightarrow C(s) = \frac{1}{s} - \frac{1}{s+\frac{1}{T}}$$

Apply inverse Laplace transform on both the sides.

$$c(t) = \left(1 - e^{-\left(\frac{t}{T}\right)}\right) u(t)$$

The **unit step response**, c(t) has both the transient and the steady state terms.

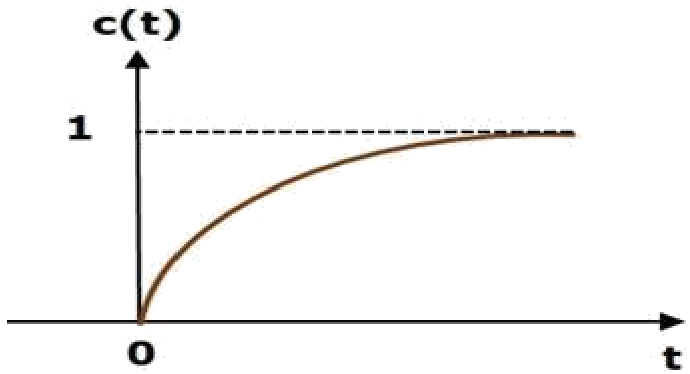
The transient term in the unit step response is -

$$c_{tr}(t) = -e^{-\left(\frac{t}{T}\right)} u(t)$$

The steady state term in the unit step response is –

$$c_{ss}(t) = u(t)$$

The following figure shows the unit step response



The value of the **unit step response, c(t)** is zero at t = 0 and for all negative values of t. It is gradually increasing from zero value and finally reaches to one in steady state. So, the steady state value depends on the magnitude of the input.

(ii) Unit impulse response:

Consider the **unit impulse signal** as an input to the first order system.

So, $r(t)=\delta(t)$

Apply Laplace transform on both the sides.

$R(s) = 1$

Consider the equation, $C(s) = \left(\frac{1}{sT+1}\right) R(s)$

Substitute, $R(s) = 1$ in the above equation.

$$C(s) = \left(\frac{1}{sT+1}\right)(1) = \frac{1}{sT+1}$$

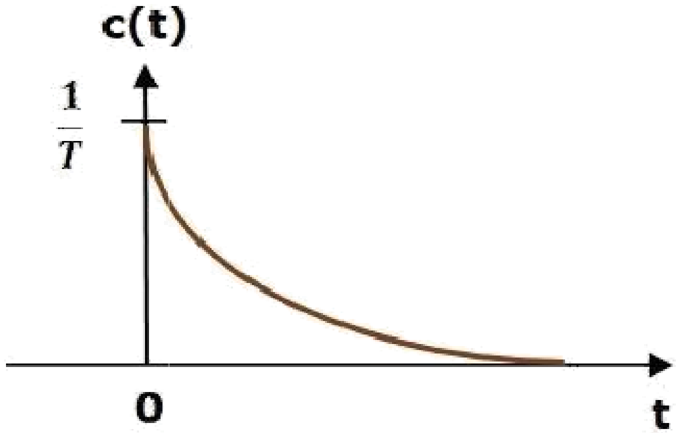
Rearrange the above equation in one of the standard forms of Laplace transforms.

$$C(s) = \frac{1}{T\left(s+\frac{1}{T}\right)} \Rightarrow C(s) = \frac{1}{T}\left(\frac{1}{s+\frac{1}{T}}\right)$$

Applying Inverse Laplace Transform on both the sides,

$$c(t) = \frac{1}{T} e^{\left(-\frac{t}{T}\right)} u(t)$$

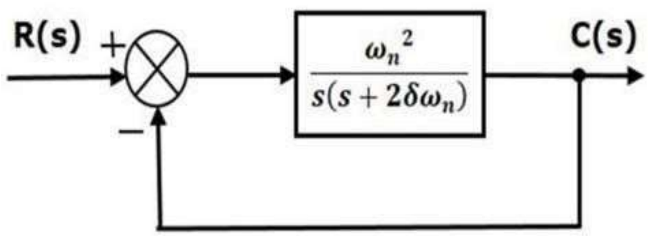
The unit impulse response is shown in the following figure.



The **unit impulse response**, c(t) is an exponential decaying signal for positive values of 't' and it is zero for negative values of 't'.

TIME RESPONSE OF 2ND ORDER SYSTEM

Consider the following block diagram of closed loop control system. Here, an open loop transfer function, ωn^2 / s(s+2δωn) is connected with a unity negative feedback.



We know that the transfer function of the closed loop control system having unity negative feedback as

$$\frac{C(s)}{R(s)} = \frac{G(s)}{1+G(s)}$$

Substitute, $G(s) = \frac{\omega_n^2}{s(s+2\delta\omega_n)}$ in the above equation.

$$\frac{C(s)}{R(s)} = \frac{\left(\frac{\omega_n^2}{s(s+2\delta\omega_n)}\right)}{1+\left(\frac{\omega_n^2}{s(s+2\delta\omega_n)}\right)} = \frac{\omega_n^2}{s^2+2\delta\omega_n s+\omega_n^2}$$

STEP RESPONSE OF 2ND ORDER SYSTEM

Consider the unit step signal as an input to the second order system. Laplace transform of the unit step signal is,

$$T(s) = \frac{C(s)}{R(s)} = \frac{\omega_n^2}{s^2 + 2\zeta\omega_n s + \omega_n^2}$$

$$Now,\ r(t) = 1\ or\ R(s) = \frac{1}{s}$$

$$\therefore C(s) = \frac{1}{s} \cdot \frac{\omega_n^2}{s^2 + 2\zeta\omega_n s + \omega_n^2}$$

$$= \frac{1}{s} \cdot \frac{\omega_n^2}{s^2 + 2\zeta\omega_n s + \zeta^2\omega_n^2 + \omega_n^2 - \zeta^2\omega_n^2}$$

$$= \frac{1}{s} \cdot \frac{\omega_n^2}{(s + \zeta\omega_n)^2 + \omega_n^2(1 - \zeta^2)}$$

$$= \frac{(s + \zeta\omega_n)^2 + \omega_n^2(1 - \zeta^2) - (s + \zeta\omega_n)^2 + \omega_n^2\zeta^2}{s\left\{(s + \zeta\omega_n)^2 + \omega_n^2(1 - \zeta^2)\right\}}$$

$$= \frac{(s + \zeta\omega_n)^2 + \omega_n^2(1 - \zeta^2) - s^2 - 2s\zeta\omega_n - \omega_n^2\zeta^2 + \omega_n^2\zeta^2}{s\left\{(s + \zeta\omega_n)^2 + \omega_n^2(1 - \zeta^2)\right\}}$$

$$= \frac{(s + \zeta\omega_n)^2 + \omega_n^2(1 - \zeta^2) - s(s + 2s\zeta\omega_n)}{s\left\{(s + \zeta\omega_n)^2 + \omega_n^2(1 - \zeta^2)\right\}}$$

$$= \frac{1}{s} - \frac{s + 2s\zeta\omega_n}{(s + \zeta\omega_n)^2 + \omega_n^2(1 - \zeta^2)}$$

$$Putting,\ \omega_d = \omega_n\sqrt{1 - \zeta^2}$$

$$= \frac{1}{s} - \frac{s + \zeta\omega_n}{(s + \zeta\omega_n)^2 + \omega_d^2} - \frac{\zeta\omega_n}{(s + \zeta\omega_n)^2 + \omega_d^2}$$

$$= \frac{1}{s} - \frac{s + \zeta\omega_n}{(s + \zeta\omega_n)^2 + \omega_d^2} - \frac{\zeta\omega_n}{\omega_d} \cdot \frac{\omega_d}{(s + \zeta\omega_n)^2 + \omega_d^2}$$

Taking the inverse Laplace Transform of above equation, we get

$$\pounds^{-1}[C(s)] = \pounds^{-1}\left[\frac{1}{s} - \frac{s + \zeta\omega_n}{(s + \zeta\omega_n)^2 + \omega_d^2} - \frac{\zeta\omega_n}{\omega_d} \cdot \frac{\omega_d}{(s + \zeta\omega_n)^2 + \omega_d^2}\right]$$

$$= \mathcal{L}^{-1}\left[\frac{1}{s}\right] - \mathcal{L}^{-1}\left[\frac{s+\zeta\omega_n}{(s+\zeta\omega_n)^2+\omega_d^2}\right] - \mathcal{L}^{-1}\left[\frac{\zeta\omega_n}{\omega_d}\cdot\frac{\omega_d}{(s+\zeta\omega_n)^2+\omega_d^2}\right]$$

$$\therefore c(t) = 1 - e^{-\zeta\omega_n t}\cdot\cos\omega_d t - \frac{\zeta\omega_n}{\omega_d}\cdot e^{-\zeta\omega_n t}\cdot\sin\omega_d t$$

$$\because \mathcal{L}^{-1}\left[\frac{1}{s}\right] = 1,\ \mathcal{L}^{-1}\left[\frac{s+\alpha}{(s+\alpha)^2+\omega^2}\right] = e^{-\alpha t}\cos\omega t,$$

$$\mathcal{L}^{-1}\left[\frac{\omega}{(s+\alpha)^2+\omega^2}\right] = e^{-\alpha t}\sin\omega t$$

The above expression of output c(t) can be rewritten as

$$c(t) = 1 - e^{-\zeta\omega_n t}\left(\cos\omega_d t + \frac{\zeta}{\sqrt{1-\zeta^2}}\cdot\sin\omega_d t\right)$$

$$= 1 - \frac{e^{-\zeta\omega_n t}}{\sqrt{1-\zeta^2}}\left(\sqrt{1-\zeta^2}\cos\omega_d t + \zeta\cdot\sin\omega_d t\right)$$

$$\left[Say,\ \zeta = \cos\phi,\ hence,\ \sqrt{1-\zeta^2} = \sin\phi\right]$$

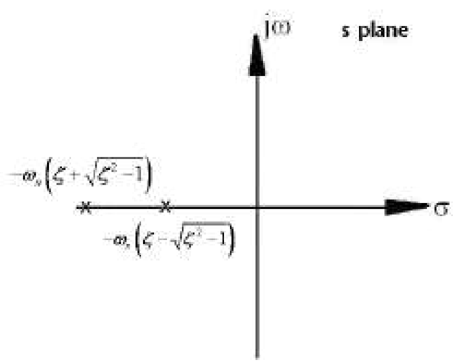
$$\therefore c(t) = 1 - \frac{e^{-\zeta\omega_n t}}{\sqrt{1-\zeta^2}}\left(\sin\phi\cos\omega_d t + \cos\phi\sin\omega_d t\right)$$

$$= 1 - \frac{e^{-\zeta\omega_n t}}{\sqrt{1-\zeta^2}}\sin\left(\omega_d t + \phi\right)$$

(a) $\zeta>1$ over damped
Poles are:

$$s_{1,2} = -\omega_n\left(\zeta \pm \sqrt{\zeta^2 - 1}\right)$$

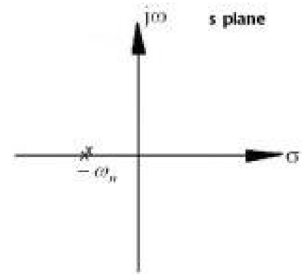
Graphically, the poles of an over damped system is shown as follows.



(b) ζ =1 critically damped
Poles are:

$$s_{1,2} = -\omega_n$$

Graphically, the poles of an critically damped system is shown as follows.



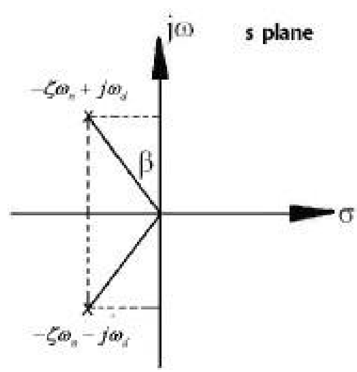
(c) ζ<1 under damped
Poles are:

$$s_{1,2} = -\omega_n\left(\zeta \pm j\sqrt{1-\zeta^2}\right)$$
$$\Rightarrow s_{1,2} = -\zeta\omega_n \pm j\omega_d$$

Where, ω_d =Damped natural frequency

$$\omega_d = \omega_n\sqrt{1-\zeta^2}$$

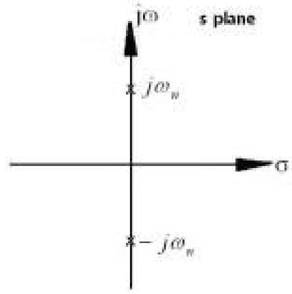
Graphically, the poles of an critically damped system is shown as follows.



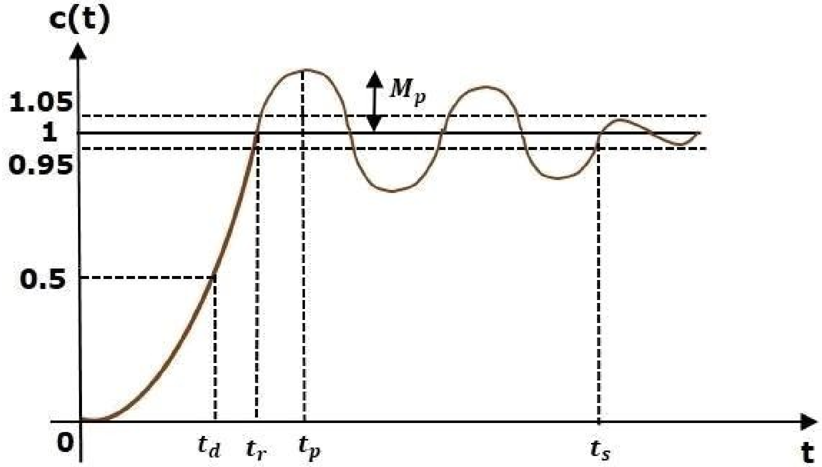
Here, $\tan\beta = \dfrac{\zeta}{\sqrt{1-\zeta^2}}$

(d) $\zeta = 0$ un-damped
Poles are:

$$s_{1,2} = -\pm j\omega_n$$



TIME RESPONSE SPECIFICATION



Delay Time

It is the time required for the response to reach **half of its final value** from the zero instant. It is denoted by t_d.

Rise Time

It is the time required for the response to rise from **0% to 100% of its final value**. This is applicable for the **under-damped systems**. For the over-damped systems, consider the duration from 10% to 90% of the final value. Rise time is denoted by **tr**.

As per definition, the magnitude of output signal at Rise times is 1. That is c(t) = 1, hence

$$1 = 1 - \frac{e^{-\zeta\omega_n t_r}}{\sqrt{1-\zeta^2}} \sin\left\{\left(\omega_n \sqrt{1-\zeta^2}\right) t_r + \phi\right\}$$

$$\Rightarrow \frac{e^{-\zeta\omega_n t_r}}{\sqrt{1-\zeta^2}} \sin\left\{\left(\omega_n \sqrt{1-\zeta^2}\right) t_r + \phi\right\} = 0$$

$$\Rightarrow \sin\left\{\left(\omega_n \sqrt{1-\zeta^2}\right) t_r + \phi\right\} = 0$$

$$\Rightarrow \left\{\left(\omega_n \sqrt{1-\zeta^2}\right) t_r + \phi\right\} = \pi$$

$$\Rightarrow t_r = \frac{\pi - \phi}{\omega_n \sqrt{1-\zeta^2}}$$

Peak Time

It is the time required for the response to reach the **peak value** for the first time. It is denoted by tp. At t=tp the first derivate of the response is zero.

As per definition at the peak time, the response curve reaches to its maximum value. Hence at that point,

$$\frac{\mathrm{d}c(t)}{\mathrm{d}t} = 0$$

$$Now,\ c(t) = 1 - \frac{e^{-\zeta\omega_n t}}{\sqrt{1-\zeta^2}} \sin\left\{\left(\omega_n\sqrt{1-\zeta^2}\right)t + \phi\right\}$$

$$\therefore \frac{\mathrm{d}c(t)}{\mathrm{d}t} = -\frac{e^{-\zeta\omega_n t}}{\sqrt{1-\zeta^2}} \cdot \omega_n\sqrt{1-\zeta^2}\cos\left\{\left(\omega_n\sqrt{1-\zeta^2}\right)t + \phi\right\}$$

$$-\frac{(-\zeta\omega_n)\,e^{-\zeta\omega_n t}}{\sqrt{1-\zeta^2}} \sin\left\{\left(\omega_n\sqrt{1-\zeta^2}\right)t + \phi\right\}$$

$$Putting,\ \frac{\mathrm{d}c(t)}{\mathrm{d}t} = 0$$

$$\therefore \frac{e^{-\zeta\omega_n t}}{\sqrt{1-\zeta^2}}\left[-\omega_n\sqrt{1-\zeta^2}\cos\left\{\left(\omega_n\sqrt{1-\zeta^2}\right)t + \phi\right\} + \zeta\omega_n \sin\left\{\left(\omega_n\sqrt{1-\zeta^2}\right)t + \phi\right\}\right]$$

$$= 0$$

$$\omega_n\sqrt{1-\zeta^2}\cos\left\{\left(\omega_n\sqrt{1-\zeta^2}\right)t + \phi\right\} = \zeta\omega_n \sin\left\{\left(\omega_n\sqrt{1-\zeta^2}\right)t + \phi\right\}$$

$$\Rightarrow \tan\left[\left(\omega_n\sqrt{1-\zeta^2}\right)t + \phi\right] = \frac{\sqrt{1-\zeta^2}}{\zeta} = \tan\phi$$

$$\therefore \left(\omega_n\sqrt{1-\zeta^2}\right)t = n\pi$$

$$Where,\ n = 1, 2, 3\ldots$$

$$t_p = \frac{\pi}{\omega_n\sqrt{1-\zeta^2}}$$

Peak Overshoot

Peak overshoot **Mp** is defined as the deviation of the response at peak time from the final value of response. It is also called the **maximum overshoot**.

Mathematically, we can write it as

$$M_p = c(t_p) - c(\infty)$$

Where,c(tp) is the peak value of the response, $c(\infty)$ is the final (steady state) value of the response.

$$c(t)_{max} = 1 - \frac{e^{-\zeta\omega_n t_p}}{\sqrt{1-\zeta^2}} \sin\left[\left(\omega_n\sqrt{1-\zeta^2}\right) t_p + \phi\right]$$

$$\Rightarrow c(t)_{max} = 1 - \frac{e^{-\zeta\omega_n \frac{\pi}{\omega_n\sqrt{1-\zeta^2}}}}{\sqrt{1-\zeta^2}} \sin\left[\left(\omega_n\sqrt{1-\zeta^2}\right)\frac{\pi}{\omega_n\sqrt{1-\zeta^2}} + \phi\right]$$

$$\Rightarrow c(t)_{max} = 1 - \frac{e^{\frac{-\zeta\pi}{\sqrt{1-\zeta^2}}}}{\sqrt{1-\zeta^2}} \sin(\pi+\phi) = 1 - \frac{e^{\frac{-\zeta\pi}{\sqrt{1-\zeta^2}}}}{\sqrt{1-\zeta^2}}(-\sin\phi)$$

$$= 1 + \frac{e^{\frac{-\zeta\pi}{\sqrt{1-\zeta^2}}}}{\sqrt{1-\zeta^2}} \sin\phi = 1 + \frac{e^{\frac{-\zeta\pi}{\sqrt{1-\zeta^2}}}}{\sqrt{1-\zeta^2}}\sqrt{1-\zeta^2} = 1 + e^{-\frac{\zeta\pi}{\sqrt{1-\zeta^2}}}$$

$$M_p = c(t)_{max} - 1 = \left(1 + e^{-\frac{\zeta\pi}{\sqrt{1-\zeta^2}}}\right) - 1$$

$$\Rightarrow M_p = e^{-\frac{\zeta\pi}{\sqrt{1-\zeta^2}}}$$

Settling time

It is the time required for the response to reach the steady state and stay within the specified tolerance bands around the final value. In general, the tolerance bands are 2% and 5%. The settling time is denoted by ts.

The settling time for 5% tolerance band is –

The settling time for 2% tolerance band is –

$$t_s = \frac{3}{\delta\omega_n} = 3\tau$$

$$t_s = \frac{4}{\delta\omega_n} = 4\tau$$

Where, τ is the time constant and is equal to 1/δωn.

- Both the settling time ts and the time constant τ are inversely proportional to the damping ratio δ.

Both the settling time ts and the time constant τ are independent of the system gain. That means even the system gain changes, the settling time ts and time constant τ will never change.

Steady state error:-

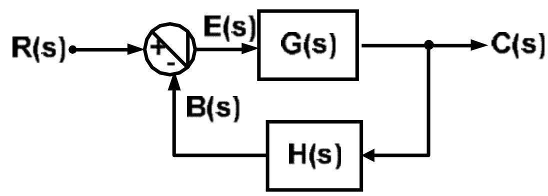
The deviation of the output of control system from desired response during steady state is known as **steady state error**. It is represented as e_{ss}. We can find steady state error using the final value theorem as follows.

$$e_{ss} = \lim_{t\to\infty} e(t) = \lim_{s\to 0} E(s)$$

Where,

E(s) is the Laplace transform of the error signal, e(t)

A simple closed-loop control system with negative feedback is shown as follows.



$E(s) = R(s) - B(s)$ --- (i)

$B(s) = C(s)\,H(s)$ ----(ii)

$C(s) = E(s)\,G(s)$---- (iii)

Applying value of B(s) of eq 2 into eq 1

$E(s) = R(s) - C(s)\,H(s)$

Applying value of C(s) of eq 3 into above eq

$$E(s) = R(s) - E(s)\,G(s)\,H(s)$$

$$\Rightarrow E(s) = \frac{R(s)}{1+G(s)H(s)}$$

Steady-state error,

$$e_{ss} = \lim_{t\to\infty} e(t) = \lim_{s\to 0} sE(s)$$

$$e_{ss} = \lim_{s\to 0} sE(s) = \lim_{s\to 0} \frac{sR(s)}{1+G(s)H(s)}$$

Therefore, steady-state error depends on two factors, i.e.

(a) type and magnitude of R(s)
(b) open-loop transfer function G(s)H(s)

Types of input and Steady-state error:

(i)Step Input:

$$R(s) = \frac{A}{s}$$

$$e_{ss}=\lim_{s\to 0}\frac{s\left(\frac{A}{s}\right)}{1+G(s)H(s)}=\lim_{s\to 0}\frac{A}{1+G(s)H(s)}$$

$$\Rightarrow e_{ss}=\frac{A}{1+\lim_{s\to 0}G(s)H(s)}=\frac{A}{1+K_P}$$

Where,

$$K_P=\lim_{s\to 0}G(s)H(s)$$

(ii) Ramp Input:

$$R(s)=\frac{A}{s^2}$$

$$e_{ss}=\lim_{s\to 0}\frac{s\left(\frac{A}{s^2}\right)}{1+G(s)H(s)}=\lim_{s\to 0}\frac{A}{s\left[1+G(s)H(s)\right]}$$

$$\Rightarrow e_{ss}=\lim_{s\to 0}\frac{A}{s+sG(s)H(s)}$$

$$\Rightarrow e_{ss}=\frac{A}{\lim_{s\to 0}sG(s)H(s)}=\frac{A}{K_V}$$

Where,

$$K_V=\lim_{s\to 0}sG(s)H(s)$$

(iii) Parabolic Input:

$$R(s)=\frac{A}{s^3}$$

$$e_{ss}=\lim_{s\to 0}\frac{s\left(\frac{A}{s^3}\right)}{1+G(s)H(s)}=\lim_{s\to 0}\frac{A}{s^2\left[1+G(s)H(s)\right]}$$

$$\Rightarrow e_{ss}=\lim_{s\to 0}\frac{A}{s^2+s^2G(s)H(s)}$$

$$\Rightarrow e_{ss}=\frac{A}{\lim_{s\to 0}s^2G(s)H(s)}=\frac{A}{K_A}$$

Where,

$$K_A=\lim_{s\to 0}s^2G(s)H(s)$$

Types of input and steady-state error are summarized as follows.

Error Constant	Equation	Steady-state error (e_{ss})
Position Error Constant (K_P)	$K_P = \lim_{s\to 0} G(s)H(s)$	$e_{ss} = \frac{A}{1+K_P}$
Velocity Error Constant (K_V)	$K_V = \lim_{s\to 0} sG(s)H(s)$	$e_{ss} = \frac{A}{K_V}$
Acceleration Error Constant (K_A)	$K_A = \lim_{s\to 0} s^2G(s)H(s)$	$e_{ss} = \frac{A}{K_A}$

STATIC ERROR COEFFICIENT METHOD

The general form of G(s)H(s) is

$$G(s)H(s) = \frac{K(1+T_1s)(1+T_2s)\dots(1+T_ns)}{s^j(1+T_as)(1+T_bs)\dots(1+T_ms)}$$

Here, j = no. of poles at origin (s = 0)

Type 0

$$G(s)H(s) = \frac{K(1+T_1s)(1+T_2s)\dots(1+T_ns)}{(1+T_as)(1+T_bs)\dots(1+T_ms)}$$

Here,

$$K_P = \lim_{s\to 0} G(s)H(s) = K$$

Therefore,

$$e_{ss} = \frac{A}{1+K}$$

Type 1

$$G(s)H(s) = \frac{K(1+T_1s)(1+T_2s)\dots(1+T_ns)}{s(1+T_as)(1+T_bs)\dots(1+T_ms)}$$

Here,

$$K_V = \lim_{s\to 0} sG(s)H(s) = K$$

Therefore,

$$e_{ss} = \frac{A}{K}$$

Type 2

$$G(s)H(s)=\frac{K(1+T_1s)(1+T_2s)...(1+T_ns)}{s^2(1+T_as)(1+T_bs)...(1+T_ms)}$$

Here,

$$K_A=\lim_{s\to 0}s^2G(s)H(s)=K$$

Therefore,

$$e_{ss}=\frac{A}{K}$$

Steady-state error and error constant for different types of input are summarized as follows.

Type	Step input		Ramp input		Parabolic input	
	K_P	e_{ss}	K_V	e_{ss}	K_A	e_{ss}
Type 0	K	$\frac{A}{1+K}$	0	∞	0	∞
Type 1	∞	0	K	$\frac{A}{K}$	0	∞
Type 2	∞	0	∞	0	K	$\frac{A}{K}$

EFFECT OF POLE AND ZERO TO TRANSFER FUNCTION

(i) Addition of a pole to the Forward Path Transfer Function:-

(a) Increases the order of the system
(b) Increases the overshoot
(c) Reduces stability
(d) Increase rise time
(e) Reduces bandwidth

(ii) Addition of a pole to the Closed-Loop Transfer function:-

(a) Increases rise time
(b) Decreases overshoot

(iii) Addition of a zero to the Closed-Loop Transfer function:-

(a) Decreases rise time
(b) Increases overshoot

(iv) Addition of a zero to the Forward path Transfer function:-

(a) Added zero far away from imaginary axis – Overshoot large & damping is very poor
(b) When zero moves to the right – Overshoot reduce & damping improves
(c) When zero moves closer to the origin – Overshoot increases & damping improves

PROPORTIONAL CONTROLLER

The proportional controller produces an output, which is proportional to error signal.

$$u(t) \propto e(t)$$

$$\Rightarrow u(t) = K_P e(t)$$

Apply Laplace transform on both the sides -

$$U(s) = K_P E(s)$$

$$\frac{U(s)}{E(s)} = K_P$$

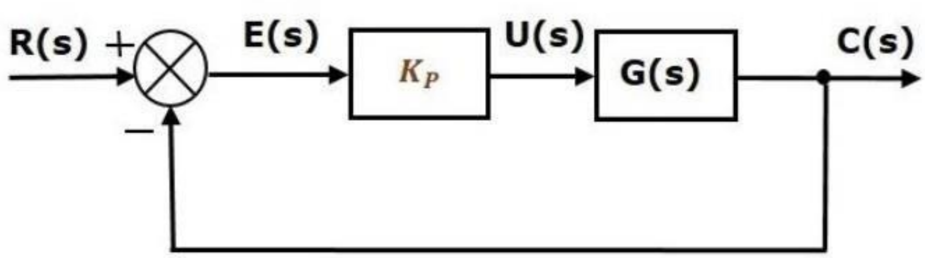
Therefore, the transfer function of the proportional controller is K_P.
Where,

U(s) is the Laplace transform of the actuating signal u(t)

E(s) is the Laplace transform of the error signal e(t)

K_P is the proportionality constant

The block diagram of the unity negative feedback closed loop control system along with the proportional controller is shown in the following figure.



DERIVATIVE CONTROLLER

The derivative controller produces an output, which is derivative of the error signal.

$$u(t) = K_D \frac{de(t)}{dt}$$

Apply Laplace transform on both sides.

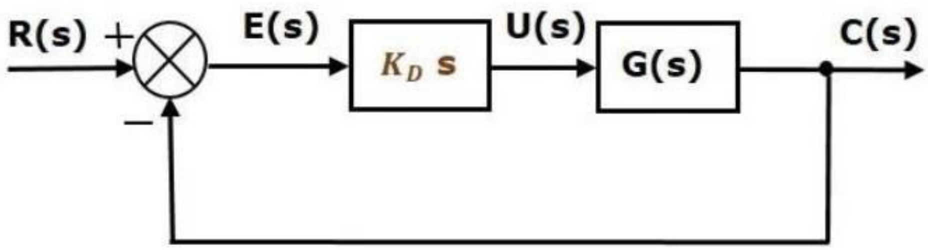
$$U(s) = K_D s E(s)$$

$$\frac{U(s)}{E(s)} = K_D s$$

Therefore, the transfer function of the derivative controller is $K_D s$

Where, K_D is the derivative constant.

The block diagram of the unity negative feedback closed loop control system along with the derivative controller is shown in the following figure.



INTEGRAL CONTROLLER

The integral controller produces an output, which is integral of the error signal.

$$u(t) = K_I \int e(t)dt$$

Apply Laplace transform on both the sides -

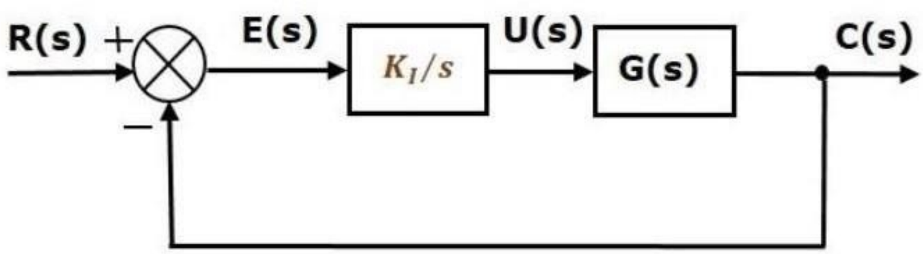
$$U(s) = \frac{K_I E(s)}{s}$$

$$\frac{U(s)}{E(s)} = \frac{K_I}{s}$$

Therefore, the transfer function of the integral controller is $\frac{K_I}{s}$.

Where, K_I is the integral constant.

The block diagram of the unity negative feedback closed loop control system along with the integral controller is shown in the following figure.



The integral controller is used to decrease the steady state error.

PROPORTIONAL DERIVATIVE (PD) CONTROLLER

The proportional derivative controller produces an output, which is the combination of the outputs of proportional and derivative controllers.

$$u(t) = K_P e(t) + K_D \frac{\mathrm{d}e(t)}{\mathrm{d}t}$$

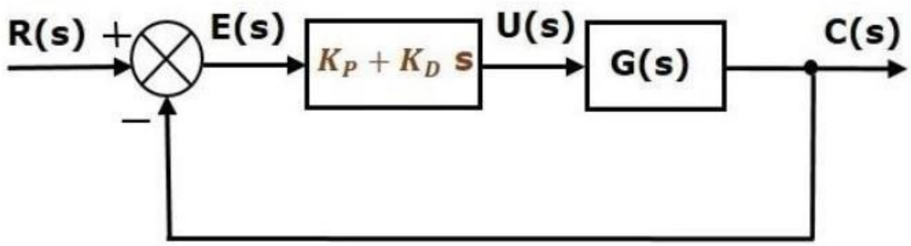
Apply Laplace transform on both sides -

$$U(s) = (K_P + K_D s)E(s)$$

$$\frac{U(s)}{E(s)} = K_P + K_D s$$

Therefore, the transfer function of the proportional derivative controller is $K_P+K_D s$.

The block diagram of the unity negative feedback closed loop control system along with the proportional derivative controller is shown in the following figure.



The proportional derivative controller is used to improve the stability of control system without affecting the steady state error.

PROPORTIONAL INTEGRAL (PI) CONTROLLER

The proportional integral controller produces an output, which is the combination of outputs of the proportional and integral controllers.

$$u(t) = K_P e(t) + K_I \int e(t)dt$$

Apply Laplace transform on both sides -

$$U(s) = \left(K_P + \frac{K_I}{s}\right) E(s)$$

$$\frac{U(s)}{E(s)} = K_P + \frac{K_I}{s}$$

Therefore, the transfer function of proportional integral controller is $K_P + \frac{K_I}{s}$.

The proportional integral controller is used to decrease the steady state error without affecting the stability of the control system.

PROPORTIONAL INTEGRAL DERIVATIVE (PID) CONTROLLER

The proportional integral derivative controller produces an output, which is the combination of the outputs of proportional, integral and derivative controllers.

$$u(t) = K_P e(t) + K_I \int e(t)dt + K_D \frac{\mathrm{d}e(t)}{\mathrm{d}t}$$

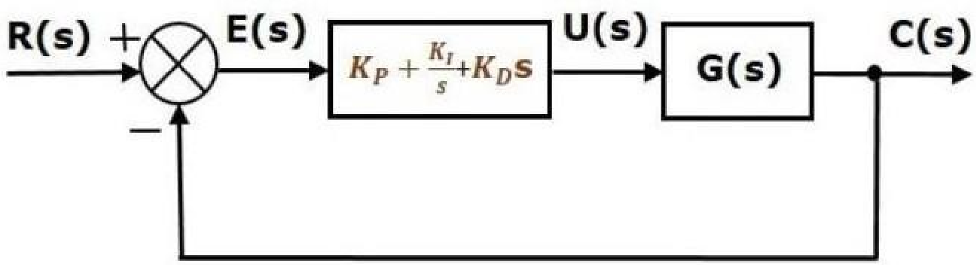
Apply Laplace transform on both sides -

$$U(s) = \left(K_P + \frac{K_I}{s} + K_D s\right) E(s)$$

$$\frac{U(s)}{E(s)} = K_P + \frac{K_I}{s} + K_D s$$

Therefore, the transfer function of the proportional integral derivative controller is $K_P + \frac{K_I}{s} + K_D s$.

The block diagram of the unity negative feedback closed loop control system along with the proportional integral derivative controller is shown in the following figure.

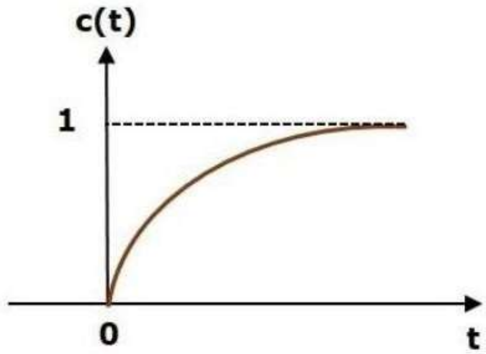


ANALYSIS OF STABILITY BY ROOT LOCUS TECHNIQUE

STABILITY

A system is said to be stable, if its output is under control. Otherwise, it is said to be unstable. A **stable system** produces a bounded output for a given bounded input.

The following figure shows the response of a stable system.



This is the response of first order control system for unit step input. This response has the values between 0 and 1. So, it is bounded output. We know that the unit step signal has the value of one for all positive values of **t** including zero. So, it is bounded input. Therefore, the first order control system is stable since both the input and the output are bounded.

TYPES OF SYSTEMS BASED ON STABILITY

We can classify the systems based on stability as follows.

- Absolutely stable system
- Conditionally stable system
- Marginally stable system

Absolutely Stable System

If the system is stable for all the range of system component values, then it is known as the **absolutely stable system**. The open loop control system is absolutely stable if all the poles of the open loop transfer function present in left half of **'s' plane**. Similarly, the closed loop control system is absolutely stable if all the poles of the closed loop transfer function present in the left half of the 's' plane.

Conditionally Stable System

If the system is stable for a certain range of system component values, then it is known as **conditionally stable system**.

Marginally Stable System

If the system is stable by producing an output signal with constant amplitude and constant frequency of oscillations for bounded input, then it is known as **marginally stable system**. The open loop control system is marginally stable if any two poles of the open loop transfer function is present on the imaginary axis. Similarly, the closed loop control system is marginally stable if any two poles of the closed loop transfer function is present on the imaginary axis.

ROUTH-HURWITZ STABILITY CRITERION

Routh-Hurwitz stability criterion is having one necessary condition and one sufficient condition for stability. If any control system doesn't satisfy the necessary condition, then we can say that the control system is unstable. But, if the control system satisfies the necessary condition, then it may or may not be stable. So, the sufficient condition is helpful for knowing whether the control system is stable or not.

NECESSARY CONDITION FOR ROUTH-HURWITZ STABILITY

The necessary condition is that the coefficients of the characteristic polynomial should be positive. This implies that all the roots of the characteristic equation should have negative real parts.

Consider the characteristic equation of the order 'n' is -

$$a_0 s^n + a_1 s^{n-1} + a_2 s^{n-2} + \ldots + a_{n-1} s^1 + a_n s^0 = 0$$

Note that, there should not be any term missing in the **n**th order characteristic equation. This means that the **n**th order characteristic equation should not have any coefficient that is of zero value.

SUFFICIENT CONDITION FOR ROUTH-HURWITZ STABILITY

The sufficient condition is that all the elements of the first column of the Routh array should have the same sign. This means that all the elements of the first column of the Routh array should be either positive or negative.

ROUTH ARRAY METHOD

If all the roots of the characteristic equation exist to the left half of the 's' plane, then the control system is stable. If at least one root of the characteristic equation exists to the right half of the 's' plane, then the control system is unstable. So, we have to find the roots of the characteristic equation to know whether the control system is stable or unstable. But, it is difficult to find the roots of the characteristic equation as order increases.

So, to overcome this problem there we have the **Routh array method**. In this method, there is no need to calculate the roots of the characteristic equation. First formulate the Routh table and find the number of the sign changes in the first column of the Routh table. The number of sign changes in the first column of the Routh table gives the number of roots of characteristic equation that exist in the right half of the 's' plane and the control system is unstable.

Follow this procedure for forming the Routh table.

- Fill the first two rows of the Routh array with the coefficients of the characteristic polynomial as mentioned in the table below. Start with the coefficient of sn and continue up to the coefficient of s_0.
- Fill the remaining rows of the Routh array with the elements as mentioned in the table below. Continue this process till you get the first column element of **row** s_0.

Note − If any row elements of the Routh table have some common factor, then you can divide the row elements with that factor for the simplification will be easy.

The following table shows the Routh array of the nth order characteristic polynomial.

$$a_0 s^n + a_1 s^{n-1} + a_2 s^{n-2} + \ldots + a_{n-1} s^1 + a_n s^0$$

s^n	a_0	a_2	a_4	a_6	$\ldots$	$\ldots$
s^{n-1}	a_1	a_3	a_5	a_7	$\ldots$	$\ldots$
s^{n-2}	$b_1 = \frac{a_1 a_2 - a_3 a_0}{a_1}$	$b_2 = \frac{a_1 a_4 - a_5 a_0}{a_1}$	$b_3 = \frac{a_1 a_6 - a_7 a_0}{a_1}$	$\ldots$	$\ldots$	$\ldots$
s^{n-3}	$c_1 = \frac{b_1 a_3 - b_2 a_1}{b_1}$	$c_2 = \frac{b_1 a_5 5 - b_3 a_1}{b_1}$	$\vdots$			
$\vdots$	$\vdots$	$\vdots$	$\vdots$			
s^1	$\vdots$	$\vdots$				
s^0	a_n					

Example

Let us find the stability of the control system having characteristic equation,

$$s^4 + 3s^3 + 3s^2 + 2s + 1 = 0$$

Step 1 − Verify the necessary condition for the Routh-Hurwitz stability.

All the coefficients of the characteristic polynomial,

$$s^4 + 3s^3 + 3s^2 + 2s + 1$$

are positive. So, the control system satisfies the necessary condition.

Step 2 − Form the Routh array for the given characteristic polynomial.

s^4	1	3	1
s^3	3	2	
s^2	$\frac{(3\times3)-(2\times1)}{3} = \frac{7}{3}$	$\frac{(3\times1)-(0\times1)}{3} = \frac{3}{3} = 1$	
s^1	$\frac{\left(\frac{7}{3}\times2\right)-(1\times3)}{\frac{7}{3}} = \frac{5}{7}$		
s^0	1		

Step 3 − Verify the sufficient condition for the Routh-Hurwitz stability.

All the elements of the first column of the Routh array are positive. There is no sign change in the first column of the Routh array. So, the control system is stable.

SPECIAL CASES OF ROUTH ARRAY

We may come across two types of situations, while forming the Routh table. It is difficult to complete the Routh table from these two situations.

The two special cases are −

- The first element of any row of the Routh's array is zero.
- All the elements of any row of the Routh's array are zero.

Let us now discuss how to overcome the difficulty in these two cases, one by one.

- **First Element of any row of the Routh's array is zero**

If any row of the Routh's array contains only the first element as zero and at least one of the remaining elements have non-zero value, then replace the first element with a small positive integer, ϵ. And then continue the process of completing the Routh's table. Now, find the number of sign changes in the first column of the Routh's table by substituting ϵ tends to zero.

- **All the Elements of any row of the Routh's array are zero**
  In this case, follow these two steps −

- Write the auxiliary equation, A(s) of the row, which is just above the row of zeros.
- Differentiate the auxiliary equation, A(s) with respect to s. fill the row of zeros with these coefficients.

ROOT LOCUS

The Root locus is the locus of the roots of the characteristic equation by varying system gain K from zero to infinity.

We know that, the characteristic equation of the closed loop control system is

$$1+G(s)H(s)=0$$

We can represent $G(s)H(s)$ as

$$G(s)H(s)=K\frac{N(s)}{D(s)}$$

Where,

- K represents the multiplying factor

- N(s) represents the numerator term having (factored) n^{th} order polynomial of 's'.
- D(s) represents the denominator term having (factored) m^{th} order polynomial of 's'.

Substitute, $G(s)H(s)$ value in the characteristic equation.

$$1 + k\frac{N(s)}{D(s)} = 0$$

$$\Rightarrow D(s) + KN(s) = 0$$

Case 1 − K = 0

If $K = 0$, then $D(s) = 0$.

That means, the closed loop poles are equal to open loop poles when K is zero.

Case 2 − K = ∞

Re-write the above characteristic equation as

$$K\left(\frac{1}{K} + \frac{N(s)}{D(s)}\right) = 0 \Rightarrow \frac{1}{K} + \frac{N(s)}{D(s)} = 0$$

Substitute, $K = \infty$ in the above equation.

$$\frac{1}{\infty} + \frac{N(s)}{D(s)} = 0 \Rightarrow \frac{N(s)}{D(s)} = 0 \Rightarrow N(s) = 0$$

If $K = \infty$, then $N(s) = 0$. It means the closed loop poles are equal to the open loop zeros when K is infinity.

From above two cases, we can conclude that the root locus branches start at open loop poles and end at open loop zeros.

ANGLE CONDITION AND MAGNITUDE CONDITION

The points on the root locus branches satisfy the angle condition. So, the angle condition is used to know whether the point exist on root locus branch or not. We can find the value of K for the points on the root locus branches by using magnitude condition. So, we can use the magnitude condition for the points, and this satisfies the angle condition.

Characteristic equation of closed loop control system is

$$1 + G(s)H(s) = 0$$

$$\Rightarrow G(s)H(s) = -1 + j0$$

The **phase angle** of $G(s)H(s)$ is

$$\angle G(s)H(s) = \tan^{-1}\left(\frac{0}{-1}\right) = (2n+1)\pi$$

The **angle condition** is the point at which the angle of the open loop transfer function is an odd multiple of 180^0

Magnitude of G(s)H(s) is –

$$|G(s)H(s)| = \sqrt{(-1)^2 + 0^2} = 1$$

The magnitude condition is that the point (which satisfied the angle condition) at which the magnitude of the open loop transfer function is one.

THE ROOT LOCUS IS A GRAPHICAL REPRESENTATION IN S-DOMAIN AND IT IS SYMMETRICAL ABOUT THE REAL AXIS. Because the open loop poles and zeros exist in the s-domain having the values either as real or as complex conjugate pairs.

RULES FOR CONSTRUCTION OF ROOT LOCUS

Follow these rules for constructing a root locus.

Rule 1 − Locate the open loop poles and zeros in the‘s’ plane.

Rule 2 − Find the number of root locus branches.

We know that the root locus branches start at the open loop poles and end at open loop zeros. So, the number of root locus branches **N** is equal to the number of finite open loop poles **P** or the number of finite open loop zeros **Z**, whichever is greater.

Mathematically, we can write the number of root locus branches **N** as

$N=P$ if $P \geq Z$
$N=Z$ if $P<Z$

Rule 3 − Identify and draw the **real axis root locus branches**.

If the angle of the open loop transfer function at a point is an odd multiple of 180^0, then that point is on the root locus. If odd number of the open loop poles and zeros exist to the left side of a point on the real axis, then that point is on the root locus branch. Therefore, the branch of points which satisfies this condition is the real axis of the root locus branch.

Rule 4 − Find the centroid and the angle of asymptotes.

- If P=Z, then all the root locus branches start at finite open loop poles and end at finite open loop zeros.
- If P>Z, then Z number of root locus branches start at finite open loop poles and end at finite open loop zeros and P−Z number of root locus branches start at finite open loop poles and end at infinite open loop zeros.
- If P<Z , then P number of root locus branches start at finite open loop poles and end at finite open loop zeros and Z−P number of root locus branches start at infinite open loop poles and end at finite open loop zeros.

So, some of the root locus branches approach infinity, when P≠Z. Asymptotes give the direction of these root locus branches. The intersection point of asymptotes on the real axis is known as **centroid**.

We can calculate the **centroid α** by using this formula,

$$\alpha = \frac{\sum Real\ part\ of\ finite\ open\ loop\ poles - \sum Real\ part\ of\ finite\ open\ loop\ zeros}{P-Z}$$

The formula for the angle of **asymptotes θ** is

$$\theta = \frac{(2q+1)180^0}{P-Z}$$

Where,

$$q = 0, 1, 2, \ldots, (P-Z)-1$$

Rule 5 − Find the intersection points of root locus branches with an imaginary axis.

We can calculate the point at which the root locus branch intersects the imaginary axis and the value of **K** at that point by using the Routh array method and special **case (ii)**.

- If all elements of any row of the Routh array are zero, then the root locus branch intersects the imaginary axis and vice-versa.
- Identify the row in such a way that if we make the first element as zero, then the elements of the entire row are zero. Find the value of **K** for this combination.
- Substitute this **K** value in the auxiliary equation. You will get the intersection point of the root locus branch with an imaginary axis.

Rule 6 − Find Break-away and Break-in points.

- If there exists a real axis root locus branch between two open loop poles, then there will be a **break-away point** in between these two open loop poles.
- If there exists a real axis root locus branch between two open loop zeros, then there will be a **break-in point** in between these two open loop zeros.

Note − Break-away and break-in points exist only on the real axis root locus branches.

Follow these steps to find break-away and break-in points.

- Write K in terms of s from the characteristic equation 1+G(s)H(s)=0.
- Differentiate K with respect to s and make it equal to zero. Substitute these values of s in the above equation.
- The values of s for which the K value is positive are the **break points**.

Rule 7 − Find the angle of departure and the angle of arrival.

The Angle of departure and the angle of arrival can be calculated at complex conjugate open loop poles and complex conjugate open loop zeros respectively.

The formula for the **angle of departure** ϕ_d is

$$\phi_d = 180^0 - \phi$$

The formula for the **angle of arrival** ϕ_a is

$$\phi_a = 180^0 + \phi$$

Where,

$$\phi = \sum \phi_P - \sum \phi_Z$$

Example

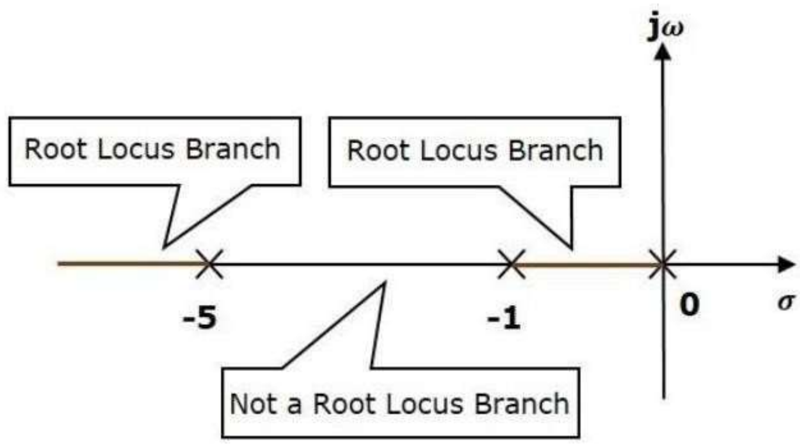
Let us now draw the root locus of the control system having open loop transfer function, $G(s)H(s) = \frac{K}{s(s+1)(s+5)}$

Step 1 − The given open loop transfer function has three poles at s = 0,

s = -1, s = -5. It doesn't have any zero. Therefore, the number of root locus branches is equal

to the number of poles of the open loop transfer function.

N=P=3

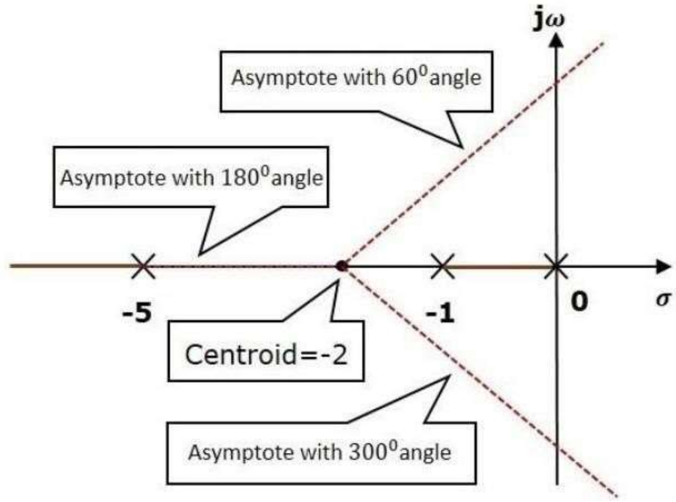


The three poles are located are shown in the above figure. The line segment between s=−1, and s=0 is one branch of root locus on real axis. And the other branch of the root locus on the real axis is the line segment to the left of s=−5.

Step 2 − We will get the values of the centroid and the angle of asymptotes by using the given formulae.

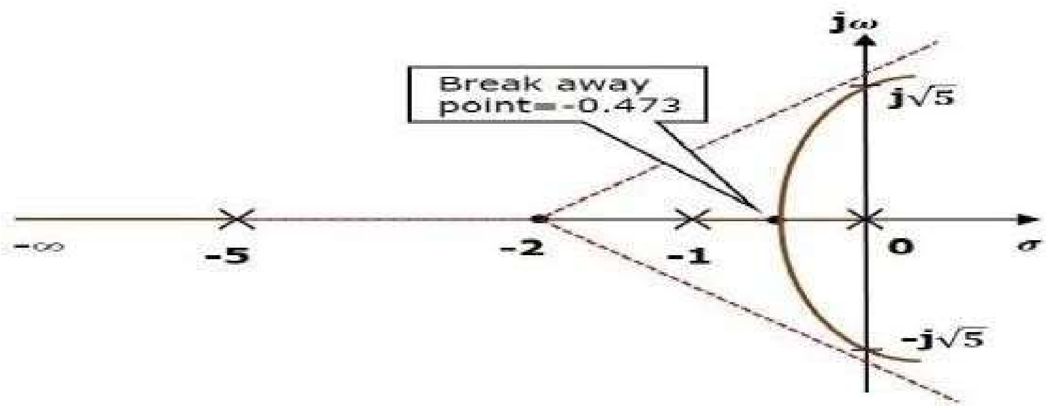
Centroid = $\frac{0-5-1}{3} = -2$

The angle of asymptotes are $\theta = 60^0, 180^0$ and 300^0.



Step 3 − Since two asymptotes have the angles of 60 and 300, two root locus branches intersect the imaginary axis. By using the Routh array method and special case(ii),
the root locus branches intersects the imaginary axis at $j\sqrt{5}$ and $-j\sqrt{5}$.

There will be one break-away point on the real axis root locus branch between the poles s =−1 and s=0. By following the procedure given for the calculation of break-away point, we will get it as s =−0.473.



EFFECTS OF ADDING OPEN LOOP POLES AND ZEROS ON ROOT LOCUS

The root locus can be shifted in‘s’ **plane** by adding the open loop poles and the open loop zeros.

- If we include a pole in the open loop transfer function, then some of root locus branches will move towards right half of ‘s’ plane. Because of this, the damping ratio δ decreases. Which implies, damped frequency ωd increases and the time domain specifications like delay time td, rise time tr and peak time tp decrease. But, it effects the system stability.

- If we include a zero in the open loop transfer function, then some of root locus branches will move towards left half of ‘s’ plane. So, it will increase the control system stability. In this case, the damping ratio δ increases. Which implies, damped frequency ωd decreases and the time domain specifications like delay time td, rise time tr and peak time tp increase.

CHAPTER-07

FREQUENCY RESPONSE ANALYSIS

FREQUENCY RESPONSE

The response of a system can be partitioned into both the transient response and the steady state response. We can find the transient response by using Fourier integrals. The steady state response of a system for an input sinusoidal signal is known as the **frequency response**. In this chapter, we will focus only on the steady state response.

If a sinusoidal signal is applied as an input to a Linear Time-Invariant (LTI) system, then it produces the steady state output, which is also a sinusoidal signal. The input and output sinusoidal signals have the same frequency, but different amplitudes and phase angles. Let the input signal be

$$r(t) = A\sin(\omega_0 t)$$

The open loop transfer function will be −

$$G(s) = G(j\omega)$$

We can represent $G(j\omega)$ in terms of magnitude and phase as shown below.

$$G(j\omega) = |G(j\omega)|\angle G(j\omega)$$

Substitute, $\omega = \omega_0$ in the above equation.

$$G(j\omega_0) = |G(j\omega_0)|\angle G(j\omega_0)$$

The output signal is

$$c(t) = A|G(j\omega_0)|\sin(\omega_0 t + \angle G(j\omega_0))$$

- The **amplitude** of the output sinusoidal signal is obtained by multiplying the amplitude of the input sinusoidal signal and the magnitude of $G(j\omega)$ at $\omega = \omega_0$.
- The **phase** of the output sinusoidal signal is obtained by adding the phase of the input sinusoidal signal and the phase of $G(j\omega)$ at $\omega = \omega_0$.

Where,

A is the amplitude of the input sinusoidal signal.

$\boldsymbol{\omega_0}$ is angular frequency of the input sinusoidal signal.

We can write, angular frequency ω0 as shown below.

$\omega_0 = 2\pi f_0$

Here, f_0 is the frequency of the input sinusoidal signal. Similarly, you can follow the same procedure for closed loop control system

FREQUENCY DOMAIN SPECIFICATIONS

The frequency domain specifications are

- **Resonant peak**
- **Resonant frequency**
- **Bandwidth**.

Consider the transfer function of the second order closed control system as

$$T(s) = \frac{C(s)}{R(s)} = \frac{\omega_n^2}{s^2 + 2\delta\omega_n s + \omega_n^2}$$

Substitute, $s = j\omega$ in the above equation.

$$T(j\omega) = \frac{\omega_n^2}{(j\omega)^2 + 2\delta\omega_n(j\omega) + \omega_n^2}$$

$$\Rightarrow T(j\omega) = \frac{\omega_n^2}{-\omega^2 + 2j\delta\omega\omega_n + \omega_n^2} = \frac{\omega_n^2}{\omega_n^2\left(1 - \frac{\omega^2}{\omega_n^2} + \frac{2j\delta\omega}{\omega_n}\right)}$$

$$\Rightarrow T(j\omega) = \frac{1}{\left(1 - \frac{\omega^2}{\omega_n^2}\right) + j\left(\frac{2\delta\omega}{\omega_n}\right)}$$

Let, $\frac{\omega}{\omega_n} = u$ Substitute this value in the above equation.

$$T(j\omega) = \frac{1}{(1 - u^2) + j(2\delta u)}$$

Magnitude of $T(j\omega)$ is -

$$M = |T(j\omega)| = \frac{1}{\sqrt{(1 - u^2)^2 + (2\delta u)^2}}$$

Phase of $T(j\omega)$ is -

$$\angle T(j\omega) = -tan^{-1}\left(\frac{2\delta u}{1 - u^2}\right)$$

Resonant Frequency

It is the frequency at which the magnitude of the frequency response has peak value for the first time. It is denoted by ω_r. At $\omega = \omega_r$, the first derivate of the magnitude of $T(j\omega)$ is zero.

Differentiate M with respect to u.

$$\frac{dM}{du} = -\frac{1}{2}\left[(1-u^2)^2 + (2\delta u)^2\right]^{\frac{-3}{2}}\left[2(1-u^2)(-2u) + 2(2\delta u)(2\delta)\right]$$

$$\Rightarrow \frac{dM}{du} = -\frac{1}{2}\left[(1-u^2)^2 + (2\delta u)^2\right]^{\frac{-3}{2}}\left[4u(u^2 - 1 + 2\delta^2)\right]$$

Substitute, $u = u_r$ and $\frac{dM}{du} == 0$ in the above equation.

$$0 = -\frac{1}{2}\left[(1-u_r^2)^2 + (2\delta u_r)^2\right]^{-\frac{3}{2}}\left[4u_r(u_r^2 - 1 + 2\delta^2)\right]$$

$$\Rightarrow 4u_r(u_r^2 - 1 + 2\delta^2) = 0$$

$$\Rightarrow u_r^2 - 1 + 2\delta^2 = 0$$

$$\Rightarrow u_r^2 = 1 - 2\delta^2$$

$$\Rightarrow u_r = \sqrt{1-2\delta^2}$$

Substitute, $u_r = \frac{\omega_r}{\omega_n}$ in the above equation.

$$\frac{\omega_r}{\omega_n} = \sqrt{1-2\delta^2}$$

$$\Rightarrow \omega_r = \omega_n\sqrt{1-2\delta^2}$$

Resonant Peak:

It is the peak (maximum) value of the magnitude of T(jω). It is denoted by Mr. At $u=u_r$, the Magnitude of T(jω) is -

$$M_r = \frac{1}{\sqrt{(1-u_r^2)^2 + (2\delta u_r)^2}}$$

Substitute, $u_r = \sqrt{1-2\delta^2}$ and $1 - u_r^2 = 2\delta^2$ in the above equation.

$$M_r = \frac{1}{\sqrt{(2\delta^2)^2 + (2\delta\sqrt{1-2\delta^2})^2}}$$

$$\Rightarrow M_r = \frac{1}{2\delta\sqrt{1-\delta^2}}$$

Resonant peak in frequency response corresponds to the peak overshoot in the time domain transient response for certain values of damping ratio δ. So, the resonant peak and peak overshoot are correlated to each other.

Bandwidth:

It is the range of frequencies over which, the magnitude of T(jω) drops to 70.7% from its zero frequency value.

At ω=0, the value of u will be zero.

Substitute, u=0 in M.

$$M = \frac{1}{\sqrt{(1-0^2)^2 + (2\delta(0))^2}} = 1$$

Therefore, the magnitude of T(jω) is one at ω=0

At 3-dB frequency, the magnitude of T(jω) will be 70.7% of magnitude of T(jω)) at ω=0

i.e., at $\omega = \omega_B, M = 0.707(1) = \frac{1}{\sqrt{2}}$

$$\Rightarrow M = \frac{1}{\sqrt{2}} = \frac{1}{\sqrt{(1-u_b^2)^2 + (2\delta u_b)^2}}$$

$$\Rightarrow 2 = (1-u_b^2)^2 + (2\delta)^2 u_b^2$$

Let, $u_b^2 = x$

$$\Rightarrow 2 = (1-x)^2 + (2\delta)^2 x$$

$$\Rightarrow x^2 + (4\delta^2 - 2)x - 1 = 0$$

$$\Rightarrow x = \frac{-(4\delta^2 - 2) \pm \sqrt{(4\delta^2 - 2)^2 + 4}}{2}$$

Consider only the positive value of x.

$$x = 1 - 2\delta^2 + \sqrt{(2\delta^2 - 1)^2 + 1}$$

$$\Rightarrow x = 1 - 2\delta^2 + \sqrt{(2 - 4\delta^2 + 4\delta^4)}$$

Substitute, $x = u_b^2 = \frac{\omega_b^2}{\omega_n^2}$

$$\frac{\omega_b^2}{\omega_n^2} = 1 - 2\delta^2 + \sqrt{(2 - 4\delta^2 + 4\delta^4)}$$

$$\Rightarrow \omega_b = \omega_n \sqrt{1 - 2\delta^2 + \sqrt{(2 - 4\delta^2 + 4\delta^4)}}$$

Bandwidth ω_b in the frequency response is inversely proportional to the rise time tr in the time domain transient response.

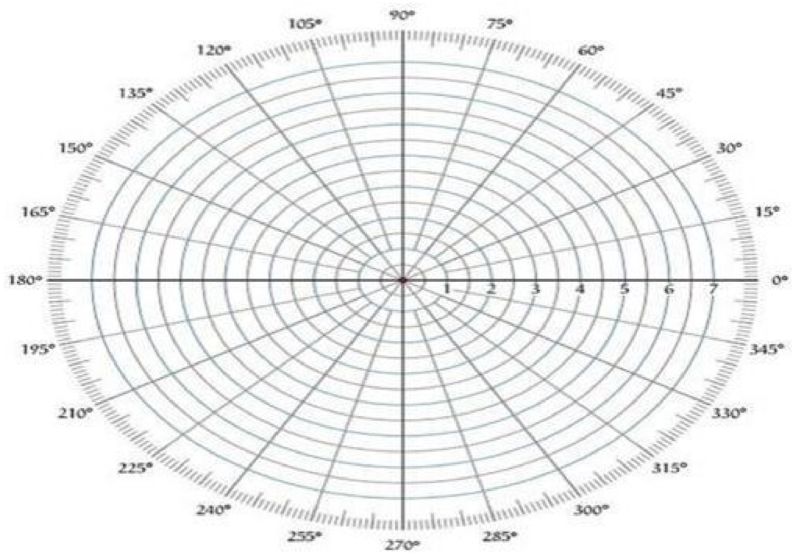
POLAR PLOTS

Polar plot is a plot which can be drawn between magnitude and phase. Here, the magnitudes are represented by normal values only.

The polar form of $G(j\omega)H(j\omega)$ is

$$G(j\omega)H(j\omega) = |G(j\omega)H(j\omega)|\angle G(j\omega)H(j\omega)$$

The **Polar plot** is a plot, which can be drawn between the magnitude and the phase angle of $G(j\omega)H(j\omega)$ by varying ω from zero to ∞. The polar graph sheet is shown in the following figure.



This graph sheet consists of concentric circles and radial lines. The **concentric circles** and the **radial lines** represent the magnitudes and phase angles respectively. These angles are represented by positive values in anti-clock wise direction. Similarly, we can represent angles with negative values in clockwise direction. For example, the angle 270^0 in anti-clock wise direction is equal to the angle -90^0 in clockwise direction.

RULES FOR DRAWING POLAR PLOTS

Follow these rules for plotting the polar plots.

- Substitute, s=jω in the open loop transfer function.
- Write the expressions for magnitude and the phase of G(jω)H(jω)
- Find the starting magnitude and the phase of G(jω)H(jω) by substituting ω=0. So, the polar plot starts with this magnitude and the phase angle.
- Find the ending magnitude and the phase of G(jω)H(jω) by substituting ω=∞ So, the polar plot ends with this magnitude and the phase angle.
- Check whether the polar plot intersects the real axis, by making the imaginary term of G(jω)H(jω) equal to zero and find the value(s) of ω.
- Check whether the polar plot intersects the imaginary axis, by making real term of G(jω)H(jω) equal to zero and find the value(s) of ω.
- For drawing polar plot more clearly, find the magnitude and phase of G(jω)H(jω) by considering the other value(s) of ω.

Example:

Consider the open loop transfer function of a closed loop control system.

$$G(s)H(s) = \frac{5}{s(s+1)(s+2)}$$

Let us draw the polar plot for this control system using the above rules.

Step 1 − Substitute, $s = j\omega$ in the open loop transfer function.

$$G(j\omega)H(j\omega) = \frac{5}{j\omega(j\omega+1)(j\omega+2)}$$

The magnitude of the open loop transfer function is

$$M = \frac{5}{\omega(\sqrt{\omega^2+1})(\sqrt{\omega^2+4})}$$

The phase angle of the open loop transfer function is

$$\phi = -90^0 - \tan^{-1}\omega - \tan^{-1}\frac{\omega}{2}$$

Frequency (rad/sec)	Magnitude	Phase angle(degrees)
0	∞	-90 or 270
∞	0	-270 or 90

So, the polar plot starts at $(\infty,-90^0)$ and ends at $(0,-270^0)$. The first and the second terms within the brackets indicate the magnitude and phase angle respectively.

Step 3 − Based on the starting and the ending polar co-ordinates, this polar plot will intersect the negative real axis. The phase angle corresponding to the negative real axis is -180^0 or 180^0. So, by equating the phase angle of the open loop transfer function to either -180^0 or 180^0, we will get the ω value as $\sqrt{2}$.

By substituting $\omega=\sqrt{2}$ in the magnitude of the open loop transfer function, we will get M=0.83. Therefore, the polar plot intersects the negative real axis when $\omega=\sqrt{2}$ and the polar coordinate is $(0.83,-180^0)$.

So, we can draw the polar plot with the above information.

BODE PLOTS

The Bode plot or the Bode diagram consists of two plots −

- Magnitude plot
- Phase plot

In both the plots, x-axis represents angular frequency (logarithmic scale). Whereas, yaxis represents the magnitude (linear scale) of open loop transfer function in the magnitude plot and the phase angle (linear scale) of the open loop transfer function in the phase plot.

The **magnitude** of the open loop transfer function in dB is -

$$M = 20\log|G(j\omega)H(j\omega)|$$

The **phase angle** of the open loop transfer function in degrees is -

$$\phi = \angle G(j\omega)H(j\omega)$$

Basic of Bode Plots:

The following table shows the slope, magnitude and the phase angle values of the terms present in the open loop transfer function. This data is useful while drawing the Bode plots.

Type of term	G(jω)H(jω)	Slope(dB/dec)	Magnitude (dB)	Phase angle(degrees)
Constant	K	0	$20\log K$	0
Zero at origin	$j\omega$	20	$20\log\omega$	90
'n' zeros at origin	$(j\omega)^n$	$20\,n$	$20\,n\log\omega$	$90\,n$
Pole at origin	$\frac{1}{j\omega}$	-20	$-20\log\omega$	$-90\ or\ 270$
'n' poles at origin	$\frac{1}{(j\omega)^n}$	$-20\,n$	$-20\,n\log\omega$	$-90\,n\ or\ 270\,n$
Simple zero	$1+j\omega r$	20	$0\ for\,\omega < \frac{1}{r}$ $20\ \log\omega r\ for\,\omega > \frac{1}{r}$	$0\ for\,\omega < \frac{1}{r}$ $90\ for\,\omega > \frac{1}{r}$
Simple pole	$\frac{1}{1+j\omega r}$	-20	$0\ for\,\omega < \frac{1}{r}$ $-20\ \log\omega r\ for\,\omega > \frac{1}{r}$	$0\ for\,\omega < \frac{1}{r}$ $-90\ or\ 270\ for\,\omega > \frac{1}{r}$
Second order derivative term	$\omega_n^2\left(1-\frac{\omega^2}{\omega_n^2}+\frac{2j\delta\omega}{\omega_n}\right)$	40	$40\ \log\ \omega_n\ for\,\omega < \omega_n$ $20\ \log\ (2\delta\omega_n^2)\ for\ \omega=\omega_n$ $40\ \log\ \omega\ for\,\omega > \omega_n$	$0\ for\,\omega < \omega_n$ $90\ for\,\omega = \omega_n$ $180\ for\,\omega > \omega_n$
Second order integral term	$\frac{1}{\omega_n^2\left(1-\frac{\omega^2}{\omega_n^2}+\frac{2j\delta\omega}{\omega_n}\right)}$	-40	$-40\ \log\ \omega_n\ for\,\omega < \omega_n$ $-20\ \log\ (2\delta\omega_n^2)\ for\ \omega=\omega_n$ $-40\ \log\ \omega\ for\,\omega > \omega_n$	$-0\ for\,\omega < \omega_n$ $-90\ for\,\omega = \omega_n$ $-180\ for\,\omega > \omega_n$

Case-1:

Consider the open loop transfer function $G(s)H(s) = K$.

Magnitude $M = 20\ \log K$ dB

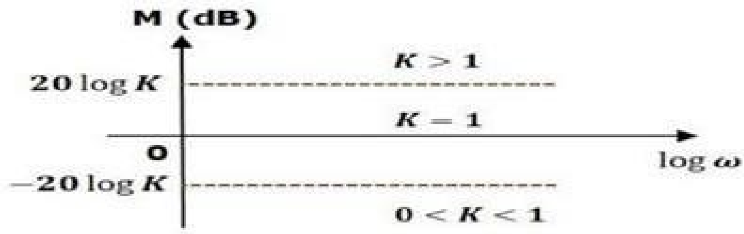
Phase angle $\phi = 0$ degrees

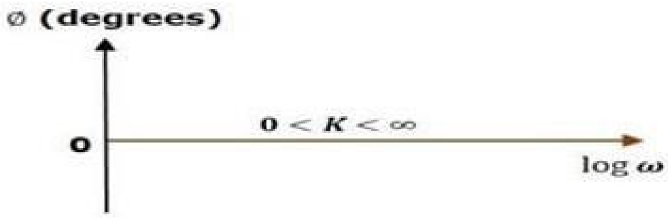
If $K = 1$, then magnitude is 0 dB.

If $K > 1$, then magnitude will be positive.

If $K < 1$, then magnitude will be negative.

The following figure shows the corresponding Bode plot.





The magnitude plot is a horizontal line, which is independent of frequency. The 0 dB line itself - is the magnitude plot when the value of K is one. For the positive values of K, the horizontal line will shift 20logK dB above the 0 dB line. For the negative values of K, the horizontal line will shift 20logK dB below the 0 dB line. The Zero degrees line itself is the phase plot for all the positive values of K.

Case-2:

Consider the open loop transfer function **G(s)H(s)= S**

Magnitude M=20logω dB

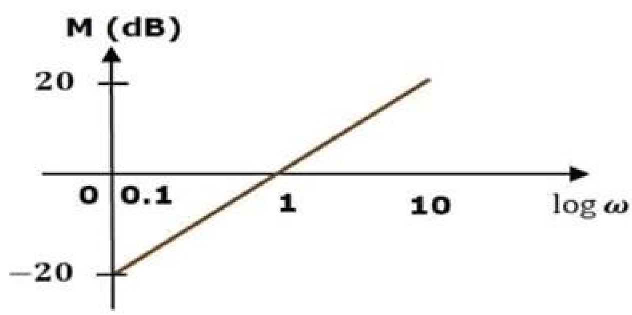
Phase angle $\phi=90^0$

At ω=0.1rad/sec, the magnitude is -20 dB.

At ω=1rad/sec, the magnitude is 0 dB.

At ω=10 rad/sec, the magnitude is 20 dB.

The following figure shows the corresponding Bode plot.



The magnitude plot is a line, which is having a slope of 20 dB/dec. This line started at ω=0.1rad/sec having a magnitude of -20 dB and it continues on the same slope. It is touching 0 dB line at ω=1 rad/sec. In this case, the phase plot is 90^0 line.

Case-3:

Consider the open loop transfer function **G(s) H(s)=1+sτ**.

Magnitude = $\sqrt{1+(s\tau)}^{2}$

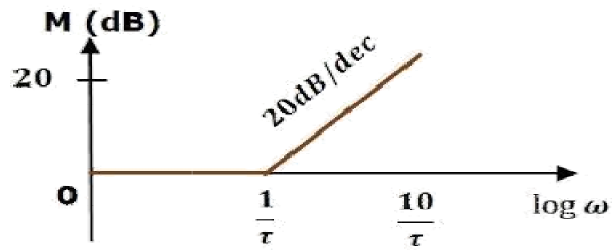
Phase angle =

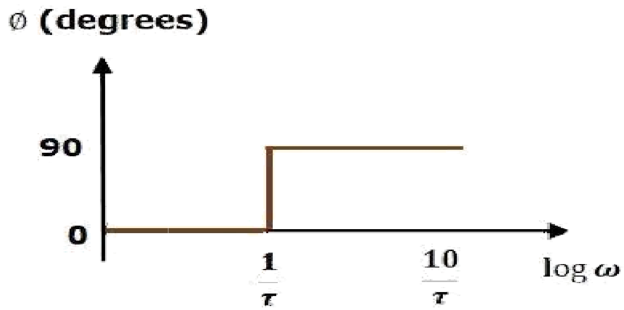
$\phi = \tan^{-1} \omega\tau$ degrees

For $\omega < \frac{1}{\tau}$, the magnitude is 0 dB and phase angle is 0 degrees.

For $\omega > \frac{1}{\tau}$, the magnitude is 20logωτ dB and phase angle is 90^0.

The following figure shows the corresponding Bode plot





The magnitude plot is having magnitude of 0 dB up to ω=1τ rad/sec. From ω=1τ rad/sec, it is having a slope of 20 dB/decade. In this case, the phase plot is having phase angle of 0 degrees up to ω=1τ rad/sec and from here, it is having phase angle of 90^0. This Bode plot is called the **asymptotic Bode plot**. As the magnitude and the phase plots are represented with straight lines, the Exact Bode plots resemble the asymptotic Bode plots. The only difference is that the Exact Bode plots will have simple curves instead of straight lines.

RULES FOR CONSTRUCTION OF BODE PLOTS:

Follow these rules while constructing a Bode plot.

- Represent the open loop transfer function in the standard time constant form.
- Substitute, s=jω in the above equation.
- Find the corner frequencies and arrange them in ascending order.
- Consider the starting frequency of the Bode plot as $1/10^{th}$ of the minimum corner frequency or 0.1 rad/sec whichever is smaller value and draw the Bode plot upto 10 times maximum corner frequency.
- Draw the magnitude plots for each term and combine these plots properly.
- Draw the phase plots for each term and combine these plots properly.

STABILITY ANALYSIS USING BODE PLOTS

From the Bode plots, we can say whether the control system is stable, marginally stable or unstable based on the values of these parameters.

- Gain cross over frequency and phase cross over frequency
- Gain margin and phase margin

Phase Cross over Frequency:

The frequency at which the phase plot is having the phase of -180^0 is known as **phase cross over frequency**. It is denoted by ω_{pc}. The unit of phase cross over frequency is **rad/sec**.

Gain Cross over Frequency:

The frequency at which the magnitude plot is having the magnitude of zero dB is known as **gain cross over frequency**. It is denoted by ω_{gc}. The unit of gain cross over frequency is **rad/sec**.

The stability of the control system based on the relation between the phase cross over frequency and the gain cross over frequency is listed below.

- If the phase cross over frequency ω_{pc} is greater than the gain cross over frequency ω_{gc}, then the control system is **stable**.
- If the phase cross over frequency ω_{pc} is equal to the gain cross over frequency ω_{gc}, then the control system is **marginally stable**.
- If the phase cross over frequency ω_{pc} is less than the gain cross over frequency ω_{gc}, then the control system is **unstable**.

Gain Margin:

Gain margin GM is equal to negative of the magnitude in dB at phase cross over frequency.

$$\mathbf{GM = -\,20\log(M_{pc})}$$

Where, Mpc is the magnitude at phase cross over frequency. The unit of gain margin (GM) is **dB**.

Phase Margin:

The formula for phase margin PM is $\mathbf{PM = 180^0 + \phi_{gc}}$

Where, ϕ_{gc} is the phase angle at gain cross over frequency. The unit of phase margin is **degrees**.

****The stability of the control system based on the relation between gain margin and phase margin is listed below.

- If both the gain margin GM and the phase margin PM are positive, then the control system is **stable**.
- If both the gain margin GM and the phase margin PM are equal to zero, then the control system is **marginally stable**.
- If the gain margin GM and / or the phase margin PM are/is negative, then the control system is **unstable**.

Example: Sketch the Bode plot for the Transfer function

$$G(s) = \frac{1000}{(1+0.1s)(1+0.001s)}$$

Determine the a) Phase Margin

b) Gain Margin

c) Stability of the System

Solution: Step-1 Put $s = j\omega$

$$G(j\omega) = \frac{1000}{(1+j0.1\omega)(1+j0.001\omega)}$$

The given transfer function is of type '0' system. Therefore the initial slope of the Bode plots 0 db/decade. The starting point is given by.

$$20 \log_{10} K = 20 \log_{10} 1000 = 60 \text{ db}$$

Corner frequencies $\omega_1 = \frac{1}{0.1} = 10$ rad/sec.

$\omega_2 = \frac{1}{0.001} = 1000$ rad/sec.

Step 2 : Mark the starting point 60 db on y-axis and draw a line of slop 0 db/decade up to first corner frequency.

Step 3 : From first corner frequency to second corner frequency draw a line with slope (0 – 20) = -20 db/decade).

Step 4 : From second corner frequency to next corner frequency (if given) draw a line having the slope -20 + (-20) = -40 db/decade.

Step 5 : The magnitude plot is complete and now draw the phase plot by calculating the phase at different frequencies (as given in table).

Step 6: From the bode plot

From the point of intersection of magnitude curve with 0 db axis draw a line on phase curve. This line cuts the phase curve at - 154°

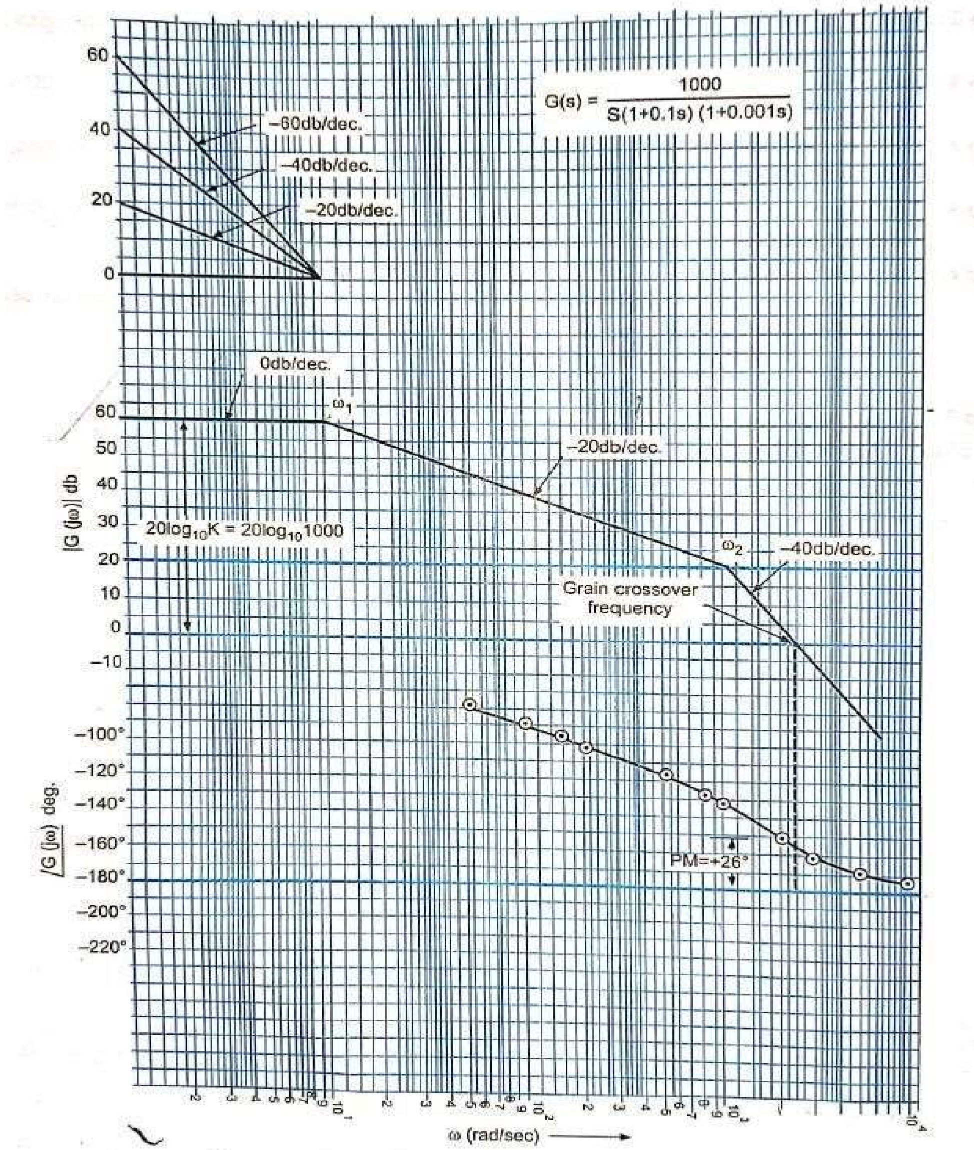
P.M = -154-(-180)

= +26°

Step 7: Gain margin G.M =∞

Since, P.M = + 26° and gain margin =∞, the system is inherently stable.

ω	$-\text{Arg}(1 + j0.1\omega)$ $-\tan^{-1}(0.1\,\omega)$	$-\text{Arg}(1 + j0.001\omega)$ $-\tan^{-1}(0.001\,\omega)$	Resultant
50	-78.6°	-2.86°	- 81.46°
100	- 84.2°	-5.7°	- 90°
150	- 86.2°	- 8.5°	- 94°
200	- 87.13°	- 11.3°	- 98°
500	- 88.85°	-26.56°	-115.4°
800	- 89.28°	- 38.65°	-127.93°
1000	-89.48°	-45°	-134.42°
2000	- 89.72°	- 63.43°	-153.15°
3000	- 89.8°	- 71.56°	-161.36°
5000	- 89.88°	- 78.69°	-168.57°
8000	- 89.92°	- 82.87°	-172.79°



CLOSED LOOP FREQUENCY RESPONSE

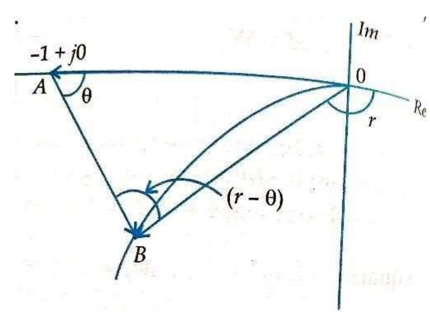
Consider transfer function $\frac{C(s)}{R(s)} = \frac{G(s)}{1 + G(s)H(s)}$

For unity feedback H(s) =1

$$\frac{C(s)}{R(s)} = \frac{G(s)}{1 + G(s)}$$

Put s= jω

$$\frac{C(j\omega)}{R(j\omega)} = \frac{G(j\omega)}{1 + G(j\omega)}$$



From figure

$$\overrightarrow{OB} = G(j\omega)$$

$$\overrightarrow{OA} = -1$$

$$\overrightarrow{AB} = \overrightarrow{OB} - \overrightarrow{OA} = G(j\omega) - (-1)$$

$$\overrightarrow{AB} = 1 + G(j\omega)$$

From above equation

$$\left|\frac{C(j\omega)}{R(j\omega)}\right| = M(\omega) = \frac{\overrightarrow{OB}}{\overrightarrow{AB}}$$

$$\angle\frac{C(j\omega)}{R(j\omega)} = \frac{\angle\overrightarrow{OB}}{\angle\overrightarrow{AB}} = \frac{r}{\theta} = r - \theta$$

$$\therefore \quad \frac{C(j\omega)}{R(j\omega)} = M(\omega)\, e^{j\,\phi(\omega)}$$

where $M(j\omega)$ is the magnitude and $\phi(\omega) = r - \theta$

Frequency response consists of 2 parts: (1) magnitude (2) phase angle. Both can be plotted against different values of ω.

CHAPTER-08

NYQUIST PLOT

NYQUIST PLOT

Nyquist plots are the continuation of polar plots for finding the stability of the closed loop control systems by varying ω from $-\infty$ to ∞. That means, Nyquist plots are used to draw the complete frequency response of the open loop transfer function.

NYQUIST STABILITY CRITERION

The Nyquist stability criterion works on the **principle of argument**. It states that if there are P poles and Z zeros are enclosed by the 's' plane closed path, then the corresponding G(s)H(s) plane must encircle the origin $P-Z$ times. So, we can write the number of encirclements N as,

$N=P-Z$

- ❖ If the enclosed's' plane closed path contains only poles, then the direction of the encirclement in the G(s)H(s) plane will be opposite to the direction of the enclosed closed path in the 's' plane.
- ❖ If the enclosed's' plane closed path contains only zeros, then the direction of the encirclement in the G(s)H(s) plane will be in the same direction as that of the enclosed closed path in the 's' plane.

Let us now apply the principle of argument to the entire right half of the's' plane by selecting it as a closed path. This selected path is called the **Nyquist** contour.

We know that the closed loop control system is stable if all the poles of the closed loop transfer function are in the left half of the's' plane. So, the poles of the closed loop transfer function are nothing but the roots of the characteristic equation. As the order of the characteristic equation increases, it is difficult to find the roots. So, let us correlate these roots of the characteristic equation as follows.

- The Poles of the characteristic equation are same as that of the poles of the open loop transfer function.
- The zeros of the characteristic equation are same as that of the poles of the closed loop transfer function.

We know that the open loop control system is stable if there is no open loop pole in the right half of the 's' plane. i.e. $P=0 \Rightarrow N=-Z$

We know that the closed loop control system is stable if there is no closed loop pole in the right half of the 's' plane. i.e. $Z=0 \Rightarrow N=P$

Nyquist stability criterion states the number of encirclements about the critical point (1+j0) must be equal to the poles of characteristic equation, which is nothing but the poles of the open loop transfer function in the right half of the 's' plane. The shift in origin to (1+j0) gives the characteristic equation plane.

RULES FOR DRAWING NYQUIST PLOTS

Follow these rules for plotting the Nyquist plots.

- Locate the poles and zeros of open loop transfer function G(s)H(s) in ‘s’ plane.
- Draw the polar plot by varying ω from zero to infinity. If pole or zero present at s = 0, then varying ω from 0+ to infinity for drawing polar plot.
- Draw the mirror image of above polar plot for values of ω ranging from $-\infty$ to zero (0^{-} if any pole or zero present at s=0).
- The number of infinite radius half circles will be equal to the number of poles or zeros at origin. The infinite radius half circle will start at the point where the mirror image of the polar plot ends. And this infinite radius half circle will end at the point where the polar plot starts.

After drawing the Nyquist plot, we can find the stability of the closed loop control system using the Nyquist stability criterion. If the critical point (-1+j0) lies outside the encirclement, then the closed loop control system is absolutely stable.

STABILITY ANALYSIS USING NYQUIST PLOTS

From the Nyquist plots, we can identify whether the control system is stable, marginally stable or unstable based on the values of these parameters.

- Gain cross over frequency and phase cross over frequency
- Gain margin and phase margin

Phase Cross over Frequency

The frequency at which the Nyquist plot intersects the negative real axis (phase angle is 180^0) is known as the **phase cross over frequency**. It is denoted by ω_{pc}.

Gain Cross over Frequency

The frequency at which the Nyquist plot is having the magnitude of one is known as the **gain cross over frequency**. It is denoted by ω_{gc}.

The stability of the control system based on the relation between phase cross over frequency and gain cross over frequency is listed below.

- If the phase cross over frequency ω_{pc} is greater than the gain cross over frequency ω_{gc}, then the control system is **stable**.
- If the phase cross over frequency ω_{pc} is equal to the gain cross over frequency ω_{gc}, then the control system is **marginally stable**.
- If phase cross over frequency ω_{pc} is less than gain cross over frequency ω_{gc}, then the control system is **unstable**.

Gain Margin

The gain margin GM is equal to the reciprocal of the magnitude of the Nyquist plot at the phase cross over frequency.

$$GM = \frac{1}{M_{pc}}$$

Where, Mpc is the magnitude in normal scale at the phase cross over frequency.

Phase Margin

The phase margin PM is equal to the sum of 180^0 and the phase angle at the gain cross over frequency.

$$PM=180^0+\phi_{gc}$$

Where, ϕ_{gc} is the phase angle at the gain cross over frequency.

The stability of the control system based on the relation between the gain margin and the phase margin is listed below.

- If the gain margin GM is greater than one and the phase margin PM is positive, then the control system is **stable**.
- If the gain margin GMs equal to one and the phase margin PM is zero degrees, then the control system is **marginally stable**.
- If the gain margin GM is less than one and / or the phase margin PM is negative, then the control system is **unstable**.

Example:- Draw the nyquist plot and assess the stability of the closed loop system whose open loop transfer function is G(s) H(s) = $\frac{K}{s(2s+1)}$

$$G(j\omega)H(j\omega) = \frac{K}{j\omega(2j\omega+1)} = \frac{K(1-j2\omega)}{j\omega(1+j2\omega)(1-j2\omega)}$$

$$= \frac{K(1-j2\omega)}{j\omega(1+4\omega^2)} = -\frac{2K}{1+4\omega^2} - j\frac{K}{\omega(1+4\omega^2)}$$

Along the segment (C_1) of the Nyquist contour on the $j\omega$-axis, s varies from $-j\infty$ to $+j\infty$.

At $\omega = -\infty$,

$$G(j\omega)H(j\omega) = -0 + j0$$

At $\omega = 0^-$,

$$G(j\omega)H(j\omega) = -2K + j\infty$$

At $\omega = 0^+$,

$$G(j\omega)H(j\omega) = -2K - j\infty$$

At $\omega = +\infty$,

$$G(j\omega)H(j\omega) = -0 - j0$$

The infinitesimally small semicircular indent around the pole at the origin of Figure represented by $s = \underset{\in \to 0}{\in e^{j\theta}}$ (θ varying from $-\pi/2°$ through 0° to $+\pi/2°$) maps into

$$\underset{\in \to 0}{\text{Lt}} \; [K/(\in e^{j\theta})(2\in e^{j\theta} + 1)] = \underset{\in \to 0}{\text{Lt}} \; K/\in e^{j\theta} = \underset{\in \to 0}{\text{Lt}} \; (K/\in)e^{-j\theta} = \infty e^{-j\theta}$$

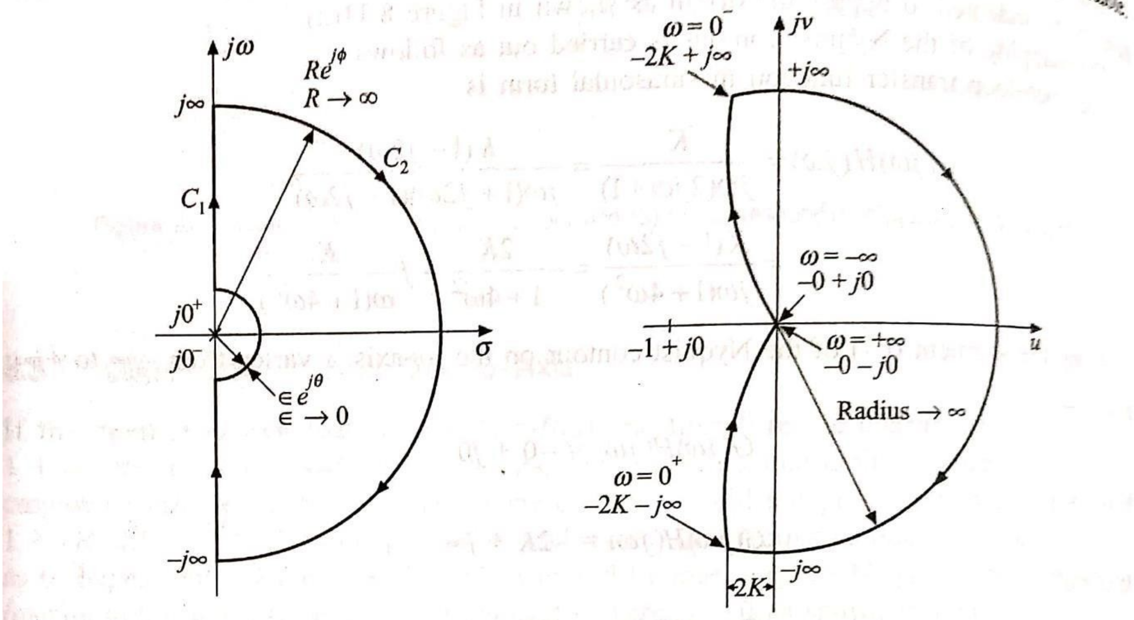
$-\theta$ varies from $+90° \to 0° \to -90°$. Thus, the infinitesimal semi-circular indent around the origin of the s-plane maps into a semicircular arc of infinite radius in $G(s)H(s)$ plane extending from +90° through 0° to −90°.

The infinite semicircular arc of the Nyquist contour (segment C_2) of Figure represented by $s = \underset{R \to \infty}{Re^{j\phi}}$ (ϕ varying from +90° through 0° to −90°) is mapped into

$$\underset{R \to \infty}{\text{Lt}} \; [K/(Re^{j\phi})(2Re^{j\phi} + 1)] = \underset{R \to \infty}{\text{Lt}} \; K/(2R^2e^{j2\phi}) = 0e^{-j2\phi}$$

that is, the origin of the $G(s)H(s)$ plane. The $G(s)H(s)$ locus thus turns at the origin with zero radius from −180° through 0° to +180°.

The given open-loop system has no poles in the right-half of the s-plane, i.e. $P = 0$. So as per the Nyquist stability criterion, for the closed-loop system to be stable, the Nyquist plot of $G(j\omega)H(j\omega)$ must not encircle the $(-1 + j0)$ point. Since the actual Nyquist plot in Figure 8.11(b) does not encircle the $(-1 + j0)$ point, the closed-loop system is always stable.



EFFECT OF ADDITION OF POLES & ZEROS TO G(s)H(s) ON THE SHAPE OF NYQUIST PLOT

a) **Addition of poles at s=0:-** It will affect the stability of the closed loop system adversely. A system that has a loop transfer function with more than one pole at s=0 is likely to be unstable or difficult to stabilize.

b) **Addition of finite non zero pole:-** It shifts the phase of nyquist plot by -90° at $\omega = \infty$. The stability is adversely affected.

c) **Addition of a Zero: -** The effect of addition of zero is to rotate the nyquist plot by 90° in the counter clockwise direction without effecting the value at $\omega = 0$. So it has the effect of reducing the overshoot & the general effect of stabilization.

CONSTANT MAGNITUDE CIRCLE (M- CIRCLE)

$$M(j\omega) = \frac{G(j\omega)}{1+G(j\omega)}$$

$$G(j\omega) = x + jy$$

$$|M(j\omega)| = \frac{\sqrt{x^2+y^2}}{\sqrt{(1+x)^2+y^2}}$$

$$M^2(1+x)^2 + M^2y^2 = x^2 + y^2$$

If M = 1, then from the above equation we obtain X = -1/2.This is the equation of a straight line parallel to the Y-axis and passing through the point (-1/2, 0).

$$x^2(1-M^2) + (1-M^2)y^2 - 2M^2x = M^2$$

Divide both the sides by $(1\text{-}M^2)$

$$x^2 + y^2 - 2\frac{M^2}{1-M^2}x = \frac{M^2}{1-M^2}$$

$\left(\frac{M^2}{1-M^2}\right)^2$ **add in both sides**

$$x^2 + y^2 - 2x\frac{M^2}{1-M^2} + \left(\frac{M^2}{1-M^2}\right)^2 = \frac{M^2}{1-M^2} + \left(\frac{M^2}{1-M^2}\right)^2$$

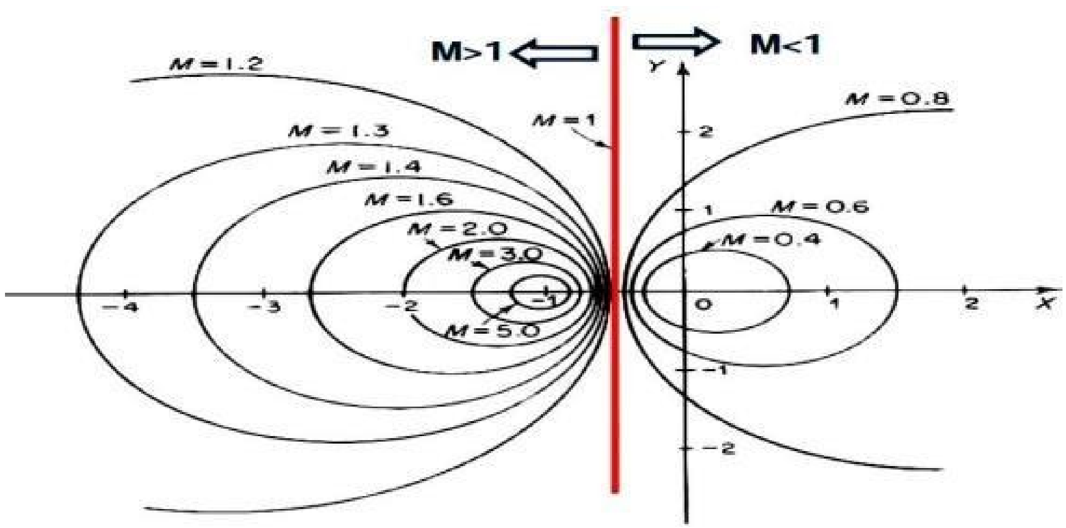
$$\left(x - \frac{M^2}{1-M^2}\right)^2 + (y-0)^2 = \frac{M^2}{1-M^2} + \frac{M^4}{(1-M^2)^2}$$

$$\left(x - \frac{M^2}{1-M^2}\right)^2 + (y-0)^2 = \frac{M^2}{(1-M^2)^2}$$

The above equation represents a family of circles with its center at $\left(\frac{M^2}{1-M^2}, 0\right)$ and radius $\left|\frac{M}{1-M^2}\right|$.

The constant M locii for different value of M. It is clear that:

i. The locii are symmetrical wrt to M= 1
ii. The M-circles for M>1 are on the left side of the line M=1 and for M<1 the constant M-circles are on the right side of the line M =1.



CONSTANT PHASE CIRCLE (N- CIRCLE):

$$\angle M = \alpha = \frac{\angle G(j\omega)}{\angle 1 + G(j\omega)}$$

$$\alpha = \tan^{-1}\frac{y}{x} - \tan^{-1}\frac{y}{1+x}$$

$$N = \tan\left(\tan^{-1}\frac{y}{x} - \tan^{-1}\frac{y}{1+x}\right)$$

$$\tan(A-B) = \frac{\tan A - \tan B}{1+\tan A \tan B}$$

Here, $\tan(\alpha)=N$

$\therefore$ $N = \dfrac{y}{x^2+x+y^2}$; or $N(x^2+x+y^2) = y$

$$x^2 + x + y^2 - \frac{y}{N} = 0$$

Add $\frac{1}{4}+\frac{1}{(2N)^2}$ to both sides, we get

$$x^2 + x + y^2 - \frac{y}{N} + \frac{1}{4} + \frac{1}{(2N)^2} = \frac{1}{4} + \frac{1}{(2N)^2}$$

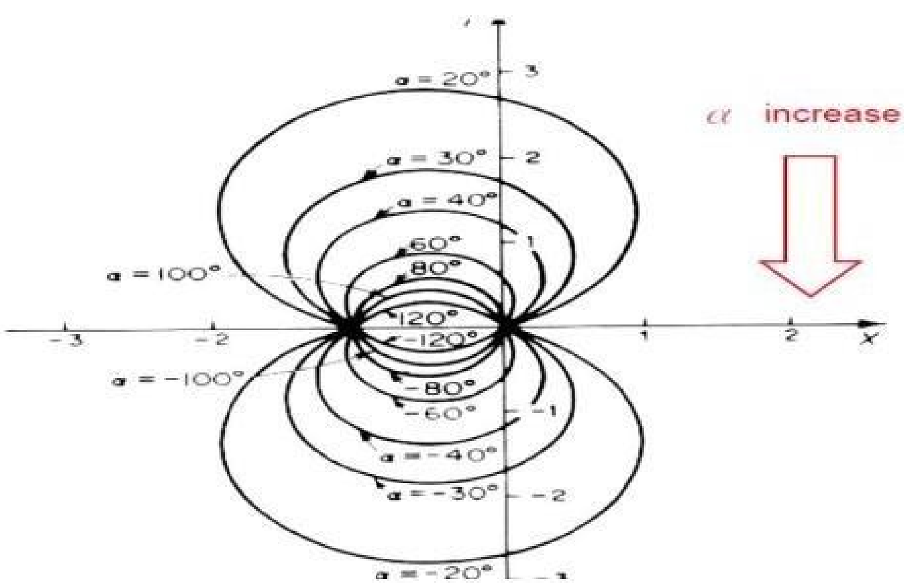
or $$(x+1/2)^2 + \left(y - \frac{1}{2N}\right)^2 = \frac{1}{4} + \frac{1}{(2N)^2}$$

The above equation represents a family of circles with its center at $(-\frac{1}{2}, \frac{1}{2N})$ and radius

$$\sqrt{\frac{1}{4}+\left(\frac{1}{2N}\right)^2}$$

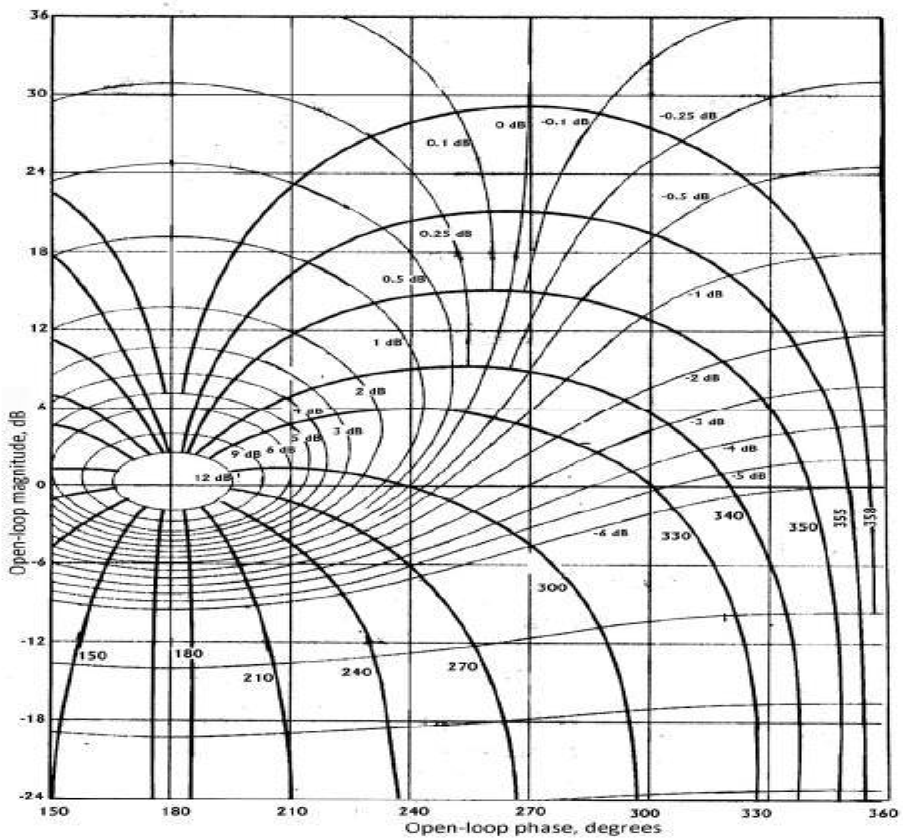
It is observed that

a) The centre is lying always at a distance x= -1/2 and y depends upon the phase shift.
b) All the circles passes through -1 as well as 0.



NICHOLS CHART

- The chart consisting of constant-magnitude loci and constant phase-angle loci in the log-magnitude versus phase diagram is called Nichols chart.
- The critical point (-1+j0) is mapped to the Nichols chart as the point (0 dB, 180degree). The Nichols chart contains curves of constant closed-loop magnitude and phase angle.
- The designer can graphically determine the phase margin, gain margin, resonant peak magnitude, resonant peak frequency, and bandwidth of the closed loops system from the plot of the open-loop locus.
- The Nichols chart is symmetric about -180 degree axis. The constant-magnitude loci and constant phase-angle loci repeat for every 360 degree, and there is a symmetry at every 180 degree. The constant-magnitude loci are centred about the critical point (0 dB, -180 degree).
- The intersection of the open-loop frequency response curve and the constant-magnitude loci and constant phase-angle loci give the values of the magnitude and the phase angle of the closed loop frequency response at each frequency point.
- If the open-loop frequency response curve does not intersect the constant-magnitude loci but is tangent to it, then the resonant peak value of the closed-loop frequency response is given by that loci. The resonant peak frequency is given by the frequency at the point of tangency.
- The phase crossover point is the point where the open-loop locus intersects the -180 degree axis, and the gain crossover point is the point where the locus intersects the 0 dB axis.
- The phase margin is the distance (measured in degrees) between the gain crossover point and the critical point (0 dB, -180 degrees).
- The gain margin is the distance (in decibels) between the phase crossover point and the critical point. The frequency at the intersection of the open-loop locus and the -3 dB locus gives the bandwidth.



Open-loop magnitude, dB
Open-loop phase, degrees
36
30
24
18
12
6
0
-6
-12
-18
-24
150
180
210
240
270
300
330
360
0 dB
-0.1 dB
-0.25 dB
0.1 dB
-0.5 dB
0.25 dB
0.5 dB
-1 dB
1 dB
-2 dB
2 dB
3 dB
-3 dB
4 dB
5 dB
6 dB
9 dB
12 dB
-4 dB
-5 dB
-6 dB
330
340
350
355
358
300
150
180
210
240
270

GANDHI ACADEMY OF TECHNOLOGY AND ENGINEERING

GOLANTHARA, BERHAMPUR

LECTURE NOTES

ON

CONTROL SYSTEM ENGINEERING

For 6th Semester

ELECTRICAL ENGINEERING

(As per Syllabus prescribed by SCTE&VT, Odisha)

Prepared By

Mr. Susanta Kumar Sahu

(Lecturer in Electrical Engineering)

Th3.CONTROL SYSTEM ENGINEERING

Name of the Course: Diploma in Electrical Engineering			
Course code:		Semester	6th
Total Period:	75	Examination	3 hrs
Theory periods:	4 P / week	Class Test:	20
Tutorial:	1 P / week	End Semester Examination:	80
Maximum marks:	100		

A. RATIONALE:

Automatic control has played a vital role in modern Engineering and Science. It has become an indispensable part of modern manufacturing and industrial process. So knowledge of automatic control system is dreadfully essential on the part of an Engineer. Basic approach to the automatic control system has been given in the subjects, so that students can enhance their knowledge in their future professional carrier.

B. OBJECTIVE:

Study of 'Control System' enhances the ability of the student on:

1. Acquire knowledge about Mathematical modeling, Block diagram algebra, signal flow graphs and control system components.
2. Ability to deal with time response analysis of various systems.
3. Finding out steady state error and error constants.
4. Acquire knowledge about the analysis of stability in Root locus technique.
5. Learning about frequency response analysis of control system.
6. To use Bode plot and Nyquist plot for judgments about stability of a system.

C. Topic wise distribution of periods:

Sl. No.	Topics	Periods
1.	Fundamental of control system	04
2.	Mathematical model of a system	04
3.	Control system components	04
4.	Block diagram algebra & signal flow graphs	08
5.	Time response analysis	10
6.	Analysis of stability by root locus technique	10
7.	Frequency response of system	10
8.	Nyquist plot	10
	Total	60

COURSE CONTENTS

1. FUNDAMENTAL OF CONTROL SYSTEM

1.1 Classification of Control system
1.2 Open loop system & Closed loop system and its comparison
1.3 Effects of Feed back
1.4 Servomechanism

2. MATHEMATICAL MODEL OF A SYSTEM

2.1 Transfer Function & Impulse response,
2.2 Properties, Advantages & Disadvantages of Transfer Function
2.3 Poles & Zeroes of transfer Function
2.4 Simple problems of transfer function of network.
2.5 Mathematical modeling of Electrical Systems(R, L, C, Analogous systems)

3. CONTROL SYSTEM COMPONENTS

3.1 Components of Control System
3.2 Gyroscope, Synchros, Tachometer, DC servomotors, Ac Servomotors.

4. BLOCK DIAGRAM ALGEBRA & SIGNAL FLOW GRAPHS

4.1 Definition: Basic Elements of Block Diagram
4.2 Canonical Form of Closed loop Systems
4.3 Rules for Block diagram reduction
4.4 Procedure for of Reduction of Block Diagram
4.5 Simple Problem for equivalent transfer function
4.6 Basic Definition in Signal Flow Graph & properties

4.7 Construction of Signal Flow graph from Block diagram
4.8 Mason's Gain formula
4.9 Simple problems in Signal flow graph for network

5. TIME RESPONSE ANALYSIS.

5 . 1 Time response of control system.
5 . 2 Standard Test signal.
- Step signal,
- Ramp Signal
- Parabolic Signal
- Impulse Signal

5 . 3 Time Response of first order system with:
- Unit step response
- Unit impulse response.

5 . 4 Time response of second order system to the unit step input.
- Time response specification.
- Derivation of expression for rise time, peak time, peak overshoot, settling time and steady stateerror. Steady state error and error constants.

5 . 5 Types of control system.[Steady state errors in Type-0, Type-1, Type-2 system]
5 . 6 Effect of adding poles and zero to transfer function.
5 . 7 Response with P, PI, PD and PID controller.

6. ANALYSIS OF STABILITY BY ROOT LOCUS TECHNIQUE.

6 . 1 Root locus concept.
6 . 2 Construction of root loci.
6 . 3 Rules for construction of the root locus.
6 . 4 Effect of adding poles and zeros to G(s) and H(s).

7. FREQUENCY RESPONSE ANALYSIS.

7 . 1 Correlation between time response and frequency response.
7 . 2 Polar plots.
7 . 3 Bode plots.
7 . 4 All pass and minimum phase system.
7 . 5 Computation of Gain margin and phase margin.
7 . 6 Log magnitude versus phase plot.
7 . 7 Closed loop frequency response.

8. NYQUIST PLOT

8.1 Principle of argument.
8.2 Nyquist stability criterion.
8.3 Niquist stability criterion applied to inverse polar plot.
8.4 Effect of addition of poles and zeros to G(S) H(S) on the shape of Niquist plot.
8.5 Assessment of relative stability.
8.6 Constant M and N circle
8.7 Nicholas chart.

CHAPTER-01

FUNDAMENTAL OF CONTROL SYSTEM

A **control system** manages commands, directs or regulates the behavior of other devices or systems using control loops. It can range from a single home heating controller using a thermostat controlling a domestic boiler to large Industrial control systems which are used for controlling processes or machines. A control system is a system, which provides the desired response by controlling the output. The following figure shows the simple block diagram of a control system.

Examples − Traffic lights control system, washing machine

Traffic lights control system is an example of control system. Here, a sequence of input signal is applied to this control system and the output is one of the three lights that will be on for some duration of time. During this time, the other two lights will be off. Based on the traffic study at a particular junction, the on and off times of the lights can be determined. Accordingly, the input signal controls the output. So, the traffic lights control system operates on time basis.

Classification of Control Systems

Based on some parameters, we can classify the control systems into the following ways.

Continuous time and Discrete-time Control Systems

- Control Systems can be classified as continuous time control systems and discrete time control systems based on the **type of the signal** used.
- In **continuous time** control systems, all the signals are continuous in time. But, in **discrete time** control systems, there exists one or more discrete time signals.

SISO and MIMO Control Systems

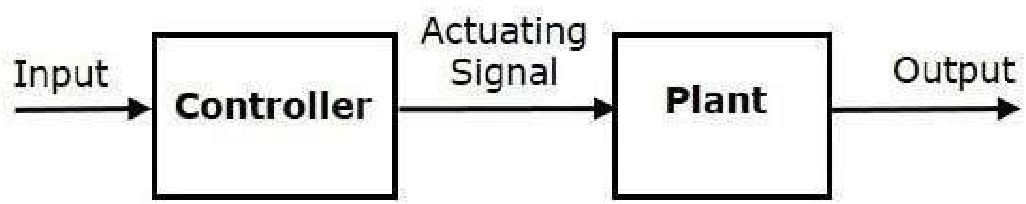
- Control Systems can be classified as SISO control systems and MIMO control systems based on the **number of inputs and outputs** present.
- **SISO** (Single Input and Single Output) control systems have one input and one output. Whereas, **MIMO** (Multiple Inputs and Multiple Outputs) control systems have more than one input and more than one output.

Open Loop and Closed Loop Control Systems

Control Systems can be classified as open loop control systems and closed loop control systems based on the **feedback path**.

In **open loop control systems**, output is not fed-back to the input. So, the control action is independent of the desired output.

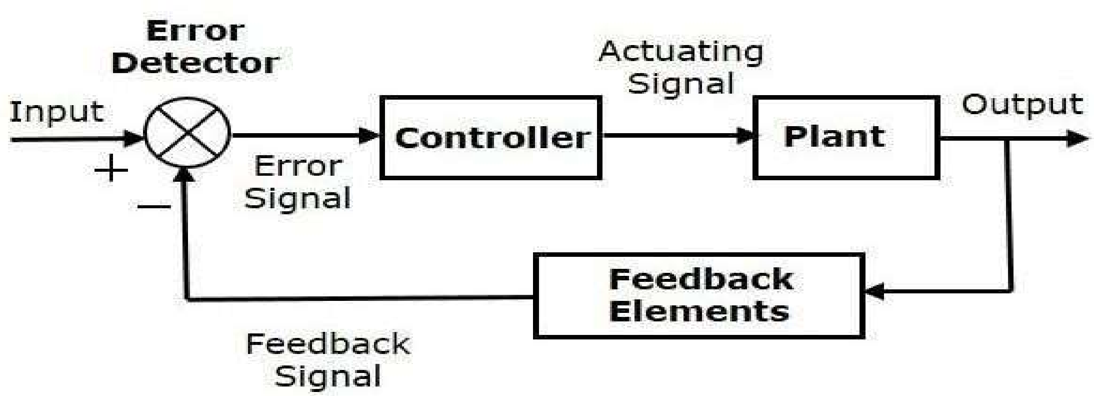
The following figure shows the block diagram of the open loop control system.



Here, an input is applied to a controller and it produces an actuating signal or controlling signal. This signal is given as an input to a plant or process which is to be controlled. So, the plant produces an output, which is controlled. The traffic lights control system which we discussed earlier is an example of an open loop control system.

In **closed loop control systems**, output is fed back to the input. So, the control action is dependent on the desired output.

The following figure shows the block diagram of negative feedback closed loop control system.



The error detector produces an error signal, which is the difference between the input and the feedback signal. This feedback signal is obtained from the block (feedback elements) by considering the output of the overall system as an input to this block. Instead of the direct input, the error signal is applied as an input to a controller.

So, the controller produces an actuating signal which controls the plant. In this combination, the output of the control system is adjusted automatically till we get the desired response. Hence, the closed loop control systems are also called the automatic control systems. Traffic lights control system having sensor at the input is an example of a closed loop control system.

The differences between the open loop and the closed loop control systems are mentioned in the following table.

Open Loop Control Systems	Closed Loop Control Systems
Control action is independent of the desired output.	Control action is dependent of the desired output.
Feedback path is not present.	Feedback path is present.
These are also called as **non-feedback control systems**.	These are also called as **feedback control systems**.
Easy to design.	Difficult to design.
These are economical.	These are costlier.
Inaccurate.	Accurate.

If either the output or some part of the output is returned to the input side and utilized as part of the system input, then it is known as **feedback**. Feedback plays an important role in order to improve the performance of the control systems. In this chapter, let us discuss the types of feedback & effects of feedback.

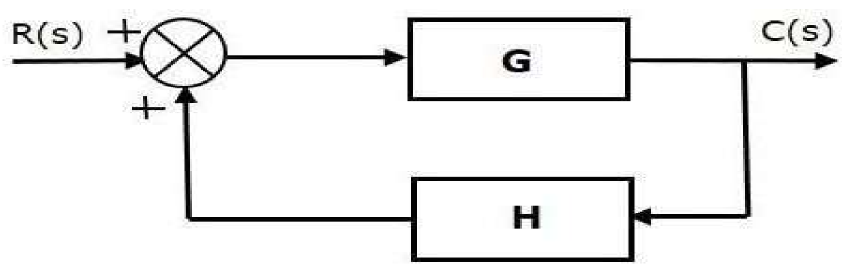
Types of Feedback

There are two types of feedback –

- Positive feedback
- Negative feedback

Positive Feedback

The positive feedback adds the reference input, R(s)R(s) and feedback output. The following figure shows the block diagram of **positive feedback control system**



he concept of transfer function will be discussed in later chapters. For the time being, consider the transfer function of positive feedback control system is,

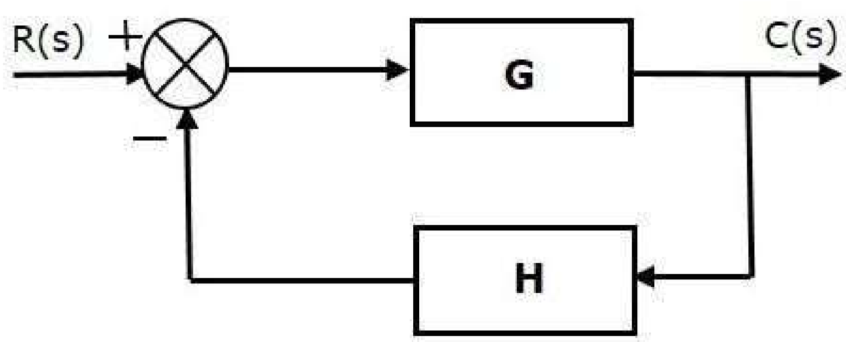
$$T = \frac{G}{1-GH} \qquad \text{(Equation 1)}$$

Where,

- **T** is the transfer function or overall gain of positive feedback control system.
- **G** is the open loop gain, which is function of frequency.
- **H** is the gain of feedback path, which is function of frequency.

Negative Feedback

Negative feedback reduces the error between the reference input, R(s)R(s) and system output. The following figure shows the block diagram of the **negative feedback control system**.



Transfer function of negative feedback control system is,

$$T = \frac{G}{1+GH} \qquad \text{(Equation 2)}$$

Where,

- **T** is the transfer function or overall gain of negative feedback control system.
- **G** is the open loop gain, which is function of frequency.
- **H** is the gain of feedback path, which is function of frequency.

The derivation of the above transfer function is present in later chapters.

Effects of Feedback

Let us now understand the effects of feedback.

Effect of Feedback on Overall Gain

- From Equation 2, we can say that the overall gain of negative feedback closed loop control system is the ratio of 'G' and (1+GH). So, the overall gain may increase or decrease depending on the value of (1+GH).
- If the value of (1+GH) is less than 1, then the overall gain increases. In this case, 'GH' value is negative because the gain of the feedback path is negative.
- If the value of (1+GH) is greater than 1, then the overall gain decreases. In this case, 'GH' value is positive because the gain of the feedback path is positive.

In general, 'G' and 'H' are functions of frequency. So, the feedback will increase the overall gain of the system in one frequency range and decrease in the other frequency range.

Effect of Feedback on Sensitivity

Sensitivity of the overall gain of negative feedback closed loop control system (**T**) to the variation in open loop gain (**G**) is defined as

$$S_G^T = \frac{\frac{\partial T}{T}}{\frac{\partial G}{G}} = \frac{Percentage\ change\ in\ T}{Percentage\ change\ in\ G} \qquad \text{(Equation 3)}$$

Where, **∂T** is the incremental change in T due to incremental change in G.

We can rewrite Equation 3 as

$$S_G^T = \frac{\partial T}{\partial G}\frac{G}{T} \qquad \text{(Equation 4)}$$

Do partial differentiation with respect to G on both sides of Equation 2.

$$\frac{\partial T}{\partial G} = \frac{\partial}{\partial G}\left(\frac{G}{1+GH}\right) = \frac{(1+GH).1-G(H)}{(1+GH)^2} = \frac{1}{(1+GH)^2} \qquad \text{(Equation 5)}$$

From Equation 2, you will get

$$\frac{G}{T} = 1 + GH \qquad \text{(Equation 6)}$$

Substitute Equation 5 and Equation 6 in Equation 4.

$$S_G^T = \frac{1}{(1+GH)^2}(1+GH) = \frac{1}{1+GH}$$

So, we got the **sensitivity** of the overall gain of closed loop control system as the reciprocal of (1+GH). So, Sensitivity may increase or decrease depending on the value of (1+GH).

- If the value of (1+GH) is less than 1, then sensitivity increases. In this case, 'GH' value is negative because the gain of feedback path is negative.
- If the value of (1+GH) is greater than 1, then sensitivity decreases. In this case, 'GH' value is positive because the gain of feedback path is positive.

In general, 'G' and 'H' are functions of frequency. So, feedback will increase the sensitivity of the system gain in one frequency range and decrease in the other frequency range. Therefore, we have to choose the values of 'GH' in such a way that the system is insensitive or less sensitive to parameter variations.

Effect of Feedback on Stability

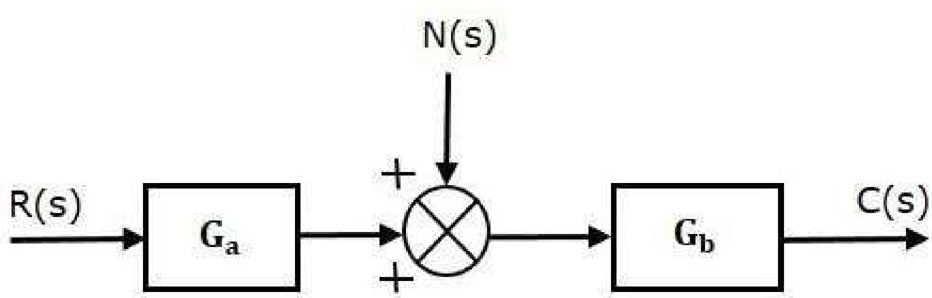
- A system is said to be stable, if its output is under control. Otherwise, it is said to be unstable.
- In Equation 2, if the denominator value is zero (i.e., GH = -1), then the output of the control system will be infinite. So, the control system becomes unstable.

Therefore, we have to properly choose the feedback in order to make the control system stable.

Effect of Feedback on Noise

To know the effect of feedback on noise, let us compare the transfer function relations with and without feedback due to noise signal alone.

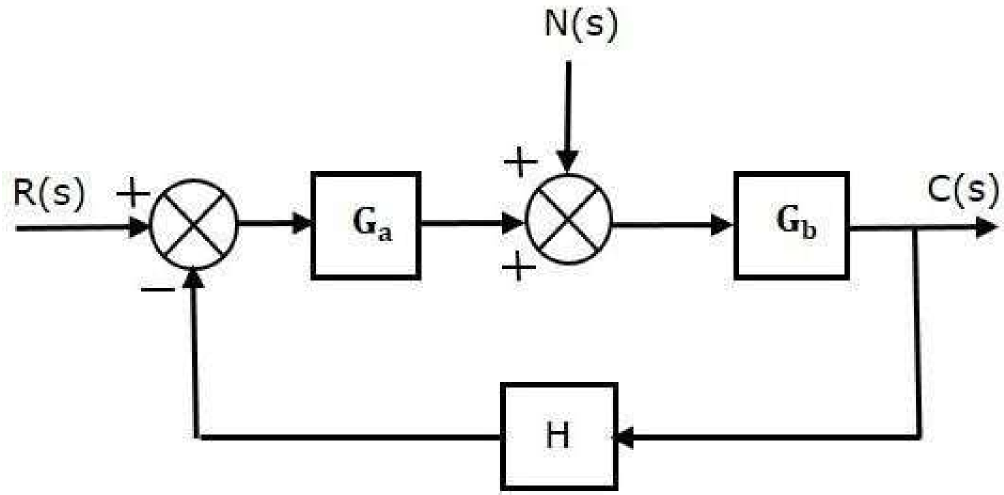
Consider an **open loop control system** with noise signal as shown below.



The **open loop transfer function** due to noise signal alone is

$$\frac{C(s)}{N(s)} = G_b \qquad \text{(Equation 7)}$$

It is obtained by making the other input $R(s)$ equal to zero.



The **closed loop transfer function** due to noise signal alone is

$$\frac{C(s)}{N(s)} = \frac{G_b}{1+G_aG_bH} \qquad \text{(Equation 8)}$$

It is obtained by making the other input $R(s)$ equal to zero.

Compare Equation 7 and Equation 8,

In the closed loop control system, the gain due to noise signal is decreased by a factor of $(1+G_aG_bH)$ provided that the term $(1+G_aG_bH)$ is greater than one.

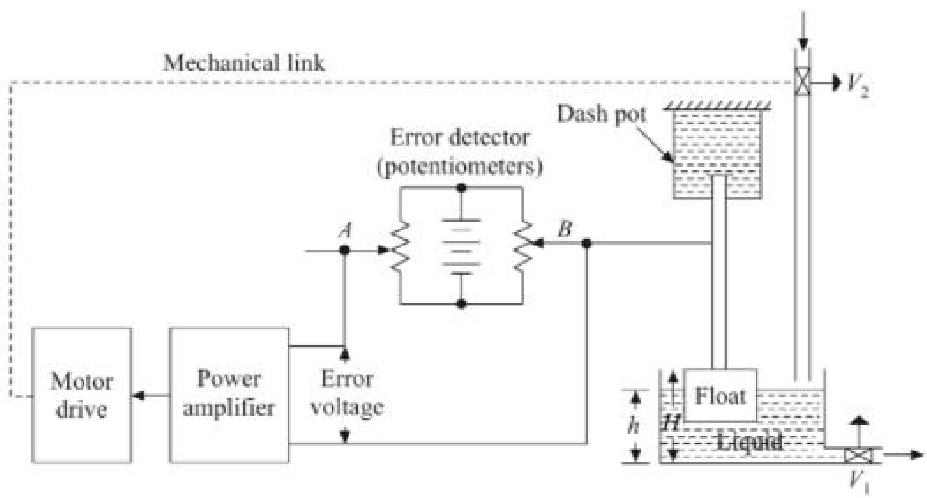
SERVOMECHANISM

In modern usage, the term servomechanism or servo is restricted to feedback controlsystems in which the controlled variable is mechanical position or time derivatives of position, e.g. velocity and acceleration. Few servo mechanisms are illustrated below.

Automatic Tank Level Control System

Figure shows an automatic tank level control system. The purpose of this system is to maintain the liquid level h (output) in the tank as close to the desired liquid level H as possible, even when the output flow rate is varied by opening the valve V_1. This has to be done by controlling the opening of the valve V_2. The potentiometer acts as an error detector. The slider arm A is positioned corresponding to the desired liquid level H (the reference input). The power amplifier and the motor drive form the control elements. The float forms the feedback path element. The valve V_2 to be controlled is the plant.

The liquid level is sensed by a float and it positions the slider arm B on the potentiometer. When the liquid level rises or falls, the potentiometer gives an error voltage proportional to the change in liquid level. The error voltage actuates the motor through a power amplifier which in turn conditions the plant (i.e. decreases or increases the opening of the valve V_2) in order to restore the desired liquid level. Thus, the control system automatically attempts to correct any deviation between the actual and desired liquid levels in the tank.



Automatic tank level control system.

CHAPTER-02

MATHEMATICAL MODEL OF A SYSTEM

The control systems can be represented with a set of mathematical equations knownas **mathematical model**. These models are useful for analysis and design of control systems.

Impulse Response and Transfer Functions

The classical way of modelling linear systems is to use transfer functions to represent input-output relations between variables. One way to define the transfer function is to use the impulse response which is defined as follows:

Impulse response: Consider that a linear time-invariant system has the input $r(t)$ and the output $c(t)$. The system can be characterized by its impulse response $g(t)$, which is defined as the output when the input is a unit impulse function $\delta(t)$. Once the impulse response of a linear system is known, the output of the system, $c(t)$, with any input $r(t)$ can be found by using the transfer function.

Transfer Function

The transfer function $G(s)$ is related to the Laplace transform of the input and the output through the following relation

$$G(s) = \frac{C(s)}{R(s)}$$

with all the initial conditions set to zero, and $C(s)$ and $R(s)$ are the Laplace transforms of $c(t)$ and $r(t)$ respectively. That is the transfer function is defined as the ratio of the Laplace transform of the output to the Laplace transform of the input with all initial conditions neglected.

The properties of the transfer function are as follows:

1. The transfer function is defined only for a linear time-invariant system. It is not defined for nonlinear systems.
2. The transfer function between an input variable and an output variable of a system is defined as the Laplace transform of the impulse response. Alternatively, the transfer function between a pair of input and output variables of a system is the ratio of the Laplace transform of the output to the Laplace transform of the input.
3. All initial conditions of the system are set to zero.
4. The transfer function is independent of the input of the system.
5. The transfer function of a continuous-data system is expressed only as a function of the complex variable s. It is not a function of the real variable time, or any other variable that is used as the independent variable. For discrete-data systems modelled by difference equations, the transfer function is a function of z when the Z transform is used

Advantages of Transfer Function

1. Once transfer function is known, output response for any type of reference input can be calculated.
2. It helps in determining the important information about the system i.e. poles', zeros, characteristic equation etc..
3. It helps in the stability analysis of the system.
4. The system differential equation can be easily obtained by replacing variable 's' by d/dt.
5. Finding inverse, the required variable can be easily expressed in the time domain. This is much more easy than to analyse the entire system in the time domain.

Disadvantages of Transfer Function

The few limitations of the transfer function approach called approach are,

i) Only applicable to linear time invariant systems.

ii) It does not provide any information concerning the physical structure of the system. From transfer function, physical nature of the system whether it is electrical, mechanical, thermal or hydraulic, cannot be judged.

iii) Effects arising due to initial conditions are totally neglected. Hence initial conditions loose their importance.

Poles & Zeroes of transfer Function

$$\text{T.F.} = \frac{P(s)}{Q(s)}$$

This can be further expressed as,

$$= \frac{a_0 s^m + a_1 s^{m-1} + a_2 s^{m-2} + \ldots + a_m}{b_0 s^n + b_1 s^{n-1} + b_2 s^{n-2} + \ldots + b_n}$$

The numerator and denominator can be factorised to get the factorised form of the transfer function as,

$$\text{T.F.} = \frac{K(s - s_a)(s - s_b) \ldots\ldots (s - s_m)}{(s - s_1)(s - s_2) \ldots\ldots (s - s_n)}$$

where K is called **system gain factor.**

values $s_1, s_2, s_3 \ldots\ldots s_n$ are called poles of the T.F.

These poles are nothing but the roots of the equation obtained by equating denominator of a T.F. to zero.

For example, let the transfer function of a system be,

$$T(s) = \frac{2(s+2)}{s(s+4)}$$

The equation obtained by equating denominator to zero is, $s(s+4) = 0$

$\therefore \quad s = 0 \quad$ and $\quad s = -4$

Similar to the poles, now if the values of 's' are substituted as s_a , s_b ......+ s_m in the numerator of a T.F., its value becomes zero.

Example *The transfer function of a system is given by,*

$$T(s) = \frac{K(s+6)}{s(s+2)(s+5)(s^2+7s+12)}$$

Determine i) Poles ii) Zeros iii) Characteristic equation and iv) Pole-zero plot in s-plane.

SOLUTION:

i) Poles are the roots of the equation obtained by equating denominator to zero i.e. roots of,

$$s(s+2)(s+5)(s^2+7s+12) = 0$$

i.e. $s(s+2)(s+5)(s+3)(s+4) = 0$

So there are 5 poles located at $s = 0, -2, -5, -3$ and -4

ii) Zeros are the roots of the equation obtained by equating numerator to zero i.e. roots of $K(s+6) = 0$

i.e. $s = -6$

There is only one zero.

iii) Characteristic equation is one, whose roots are the poles of the transfer function. So it is,

$$s(s+2)(s+5)(s^2+7s+12) = 0$$

i.e. $\quad s(s^2+7s+10)(s^2+7s+12) = 0$

i.e $s^5 + 14s^4 + 71s^3 + 154s^2 + 120s = 0$

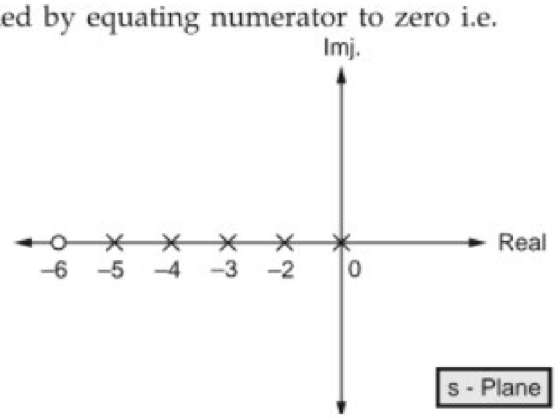


Fig.

iv) Pole-zero plot

This is shown in the Fig.

ANALOGOUS SYSTEMS

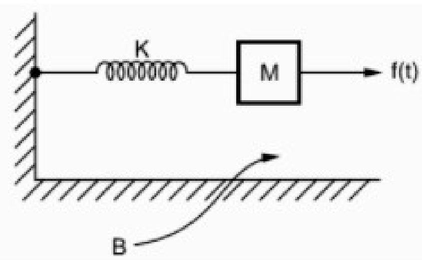
Comparing equations for the mechanical translational system or for the mechanical rotational system and for the series electrical system, it is seen that they are of identical form. Such systems whose differential equations are of identical form are called *analogous systems*. The force F (torque T) and voltage e are the analogous variables here. This is called the force (torque)-voltage analogy. A list of analogous variables in this analogy is given in Table

Table Analogous quantities in force (torque)-voltage analogy

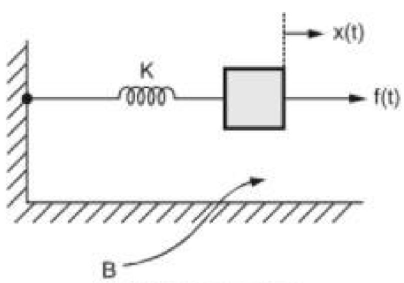
Mechanical translational system	*Mechanical rotational system*	*Electrical system*
Force F	Torque T	Voltage e
Mass M	Moment of inertia J	Inductance L
Viscous friction coefficient f	Viscous friction coefficient f	Resistance R
Spring stiffness K	Torsional spring stiffness K	Reciprocal of capacitance $1/C$
Displacement x	Angular displacement θ	Charge q
Velocity v	Angular velocity ω	Current i

Example *For the physical system shown draw its equivalent system and write equilibrium equations. Hence draw its electrical analogous circuits based on*

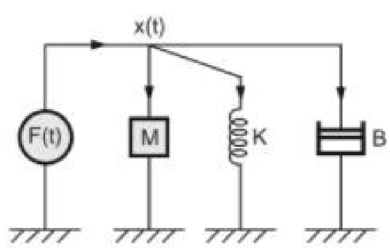
i) Force -Voltage ii) Force - Current method.



Solution : Mass \`M' will displace by amount \`x' and as spring is connected to fixed support and friction \`B' is also with respect to fixed support, both K and B will be under influence of \`x' only. Now its equivalent system will contain one node and as all elements are under influence of x(t) alone, must be connected in parallel under that node.



(a) Physical system



(b) Equivalent mechanical system

The equilibrium equation will be,

$$f(t) = M \frac{d^2 x(t)}{dt^2} + K x(t) + B \frac{dx(t)}{dt}$$

Taking Laplace, $F(s) = Ms^2 X(s) + K X(s) + Bs X(s) = X(s)\left[Ms^2 + K + Bs\right]$ ... (1)

i) Force-Voltage Method : Use the analogous terms as,

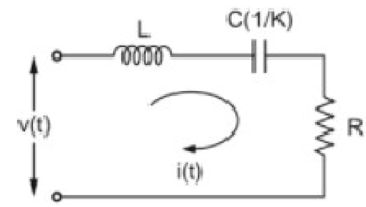
$$M \to L,\ B \to R,\ K \to \frac{1}{C},\ x \to q,\ \frac{dx}{dt} \to \frac{dq}{dt} \to i,\ x \to \int idt,\ \frac{d^2x}{dt^2} = \frac{di}{dt}$$

All quantities are expressed in terms of current i.

$$v(t) = L\frac{di}{dt} + \frac{1}{C}\int idt + Ri \quad \text{... (2)}$$

$$\therefore \quad V(s) = sL\, I(s) + \frac{I(s)}{sC} + I(s)R \quad \text{... (3)}$$

Simulate using loop method :



Analogous to K is a capacitor C but its value is proportional to 1/K hence it is indicated by writing (1/K) in the bracket near C. This is shown in the Fig.

ii) Force-Current Method : Use the analogous terms as,

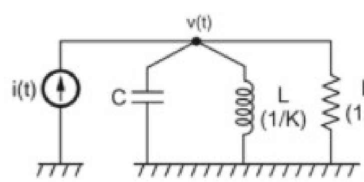
$$M \to C,\ B \to \frac{1}{R},\ K \to \frac{1}{L},\ x \to \phi,\ \frac{dx}{dt} \to \frac{d\phi}{dt} \to v(t),\ x \to \int v(t)dt,\ \frac{d^2x}{dt^2} = \frac{dv(t)}{dt}$$

All quantities are expressed in terms of voltage v.

$$\therefore \quad i(t) = C\frac{dv(t)}{dt} + \frac{1}{R}v(t) + \frac{1}{L}\int v(t)dt \quad \text{... (4)}$$

$$\therefore \quad I(s) = sCV(s) + \frac{1}{R}V(s) + \frac{1}{sL}V(s) \quad \text{... (5)}$$

Simulate using node method :



Analogous to K is an inductor L while to B is a resistor R. But their values are proportional to reciprocals of K and B respectively. This is indicated by writing (1/K) and (1/B) in the brackets near L and R respectively in the Fig.

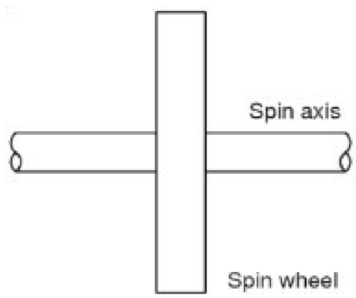
Gyroscope

Gyroscope is used in the navigation system and as gyroscope works on the inertia principle it is called inertial navigation system.

CONSTRUCTION OF THE GYROSCOPE

The Gyroscope works on the inertial principle. Inertia of the mechanical system is similar to inductance of the electrical system. In electrical system, the change of current or voltage is opposed by the inductance of the system. In mechanical system the change of position is opposed by the inertia. According to Newton's law of inertia, if any object is stationary, it will remain stationary unless acted upon by a force. If it is in motion, it will remain in motion. In linear motion, force is required to overcome the inertia given by the mass. In case of angular motion, torque is required to impart change of angular motion overcoming rotational moment of inertia.

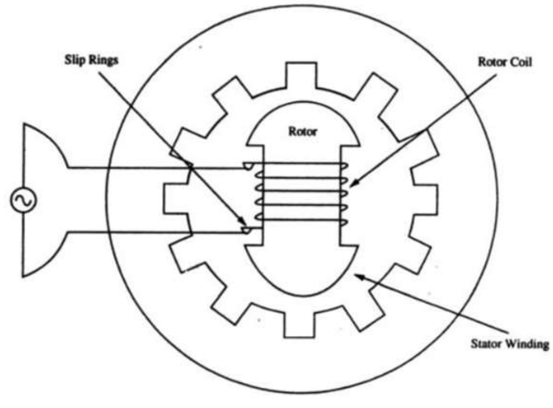
In case of gyroscope, this inertia comes from the kinetic energy stored and momentum attained by the rotating element. For this purpose the gyroscope has a spin wheel. It has a small thickness and large diameter, because the rotational moment of inertia of the disc is proportional to the forth power of the radius. To allow the rotation of a spin wheel it has an axis of rotation provided by a shaft on which the spin wheel is mounted as shown in Figure



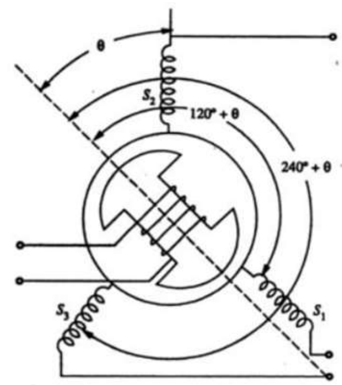
Synchros

A synchro (selsyn or autosyn) is an electromagnetic transducer that is used to convert an angular shaft position into an electric signal. The basic element of the synchro is a synchro transmitter whose construction is very similar to that of a three-phase alternator. The stator which is stationary is of laminated silicon steel and is slotted to accommodate a balanced three-phase winding which is of concentric coil type in which three identical coils are placed in the stator with their axis 120° apart and is *Y*-connected. The rotor is of dumbbell construction and is wound with a concentric coil. Through the slip rings, an a.c. voltage is applied to the rotor winding.

The constructional features and schematic diagram of a synchro transmitter are shown in Fig.



Constructional features of synchro transmitter.



Schematic diagram of synchro transmitter.

Let v_{s_1}, v_{s_2} and v_{s_3} be the voltages induced in the stator coils S_1, S_2 and S_3 with respect to the neutral respectively. Then, for the rotor position of the synchro transmitter shown in Fig. where the rotor axis makes an angle θ with the axis of the stator coil S_2

$$v_{s_1} = KV_r \sin \omega_c t \,.\, \cos(\theta + 120°)$$

$$v_{s_2} = KV_r \sin \omega_c t \,.\, \cos \theta$$

$$v_{s_3} = KV_r \sin \omega_c t \,.\, \cos(\theta + 240°)$$

The three terminal voltages of the stator are

$$v_{s_1 s_2} = v_{s_1} - v_{s_2} = \sqrt{3}\, K\, V_r \sin(\theta + 240°) \sin \omega_c t$$

$$v_{s_2 s_3} = v_{s_2} - v_{s_3} = \sqrt{3}\, KV_r \sin(\theta + 120°) \sin \omega_c t$$

$$v_{s_3 s_1} = v_{s_3} - v_{s_1} = \sqrt{3}\, KV_r \sin \theta \,.\, \sin \omega_c t$$

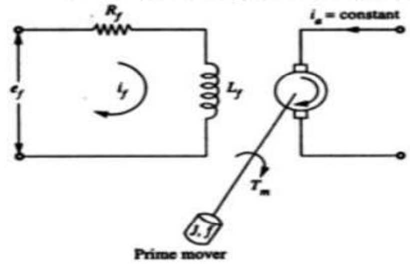
DC servomotors

While deriving transfer functions, the system is approximated by a linear lumped constant parameters model by making suitable assumptions.

As an example to illustrate this, let us derive the transfer function of a d.c. servomotor. In servo applications, a d.c. motor is expected to produce very high acceleration from standstill. The requirements of the d.c. motor are low inertia and high starting torque. A low value of inertia can be attained by reducing the armature diameter in order that the desired output power is achieved. A d.c. servomotor is an ordinary d.c. motor but for some minor changes. The two different modes in which the d.c. motor can be operated are (*i*) field control mode and (*ii*) armature control mode.

(*a*) Transfer Function of a Field Controlled D.C. Motor

A field controlled D.C. motor is shown in Fig.



Field Controlled D.C. Motor.

Let R_f and L_f be the resistance and inductance of the field circuit with excitation voltage e_f and current i_f. Let T_m be the torque developed by the motor in N-m. Let 'J' be the moment of inertia in kg-m^2, 'f' be the coefficient of friction in N-m/(rad/sec) and 'θ' be the angular displacement of motor shaft in radians.

Applying Kirchhoff's law for the field circuit

$$e_f(t) = R_f i_f(t) + L_f \frac{di_f(t)}{dt}$$

Taking Laplace Transform of equation (5.19), we have,

$$E_f(s) = (R_f + sL_f)\ I_f(s)$$

$$I_f(s) = \frac{E_f(s)}{(R_f + sL_f)}$$

The torque developed is proportional to the field current, since the armature current is constant.

$$T_m(t) = K_t i_f(t)$$

where K_t is a constant.

Taking Laplace Transform of equation (5.21),

$$T_m(s) = K_t \,.\, I_f(s)$$

The torque equation is given by

$$T_m(t) = J\frac{d^2\theta}{dt^2} + f\frac{d\theta}{dt}$$

Taking Laplace Transform,

$$T_m(s) = (Js^2 + fs)\ \theta(s)$$

$$(Js^2 + fs)\ \theta(s) = K_t \,.\, I_f(s)$$

$$s(Js + f)\ \theta(s) = K_t \cdot \frac{E_f(s)}{(R_f + sL_f)}$$

The transfer function is given by

$$G(s) = \frac{\theta(s)}{E_f(s)}$$

$$= \frac{K_t}{s(Js+f)\ (R_f+sL_f)}$$

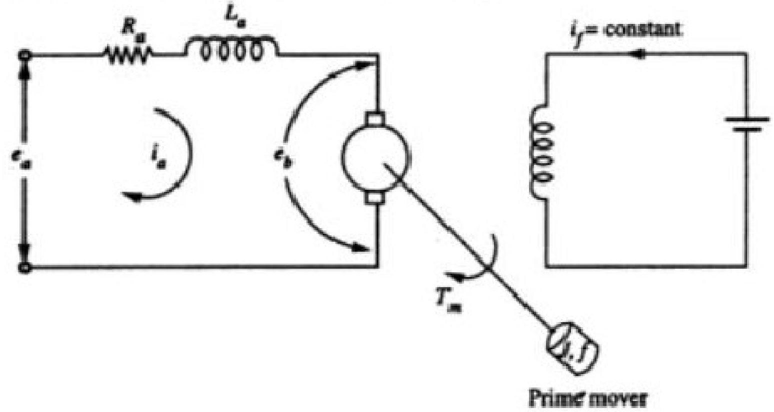
$$= \frac{K_t}{fs\left[1+s\left(\frac{J}{f}\right)\right] R_f\left[1 + s\left(\frac{L_f}{R_f}\right)\right]}$$

$$G(s) = \frac{K_t'}{s(1 + \tau_{me})\ (1 + s\,\tau_f)}$$

where
$$K_t' = \frac{K_t}{R_f - f}$$
$$\tau_{me} = \frac{J}{f} = \text{Mechanical time constant}$$
$$\tau_f = \frac{L_f}{R_f} = \text{field time constant}$$

(*b*) Transfer Function of an Armature Controlled D.C. Motor

A D.C. armature controlled motor is shown in Fig.



Armature Controlled D.C. Motor.

Let R_a and L_a be the resistance and inductance of the armature circuit with an applied voltage of e_a. Let e_b be the back e.m.f. in the armature circuit. Let i_f be the constant field current through the field windings. Let T_m be the torque developed in N-m, 'J' be the moment of inertia in kg-m^2, 'f' be the friction coefficient in $\frac{\text{N–m}}{\text{rad/sec}}$ and θ, the angular displacement in radians.

Applying Kirchhoff's voltage law to the armature circuit,

$$e_a(t) = R_a\, i_a(t) + L_a \,.\, \frac{di_a(t)}{dt} + e_b(t)$$

Taking Laplace Transform of equation , we have,

$$E_a(s) = [R_a + sL_a]\ I_a(s) + E_b(s)$$

$$I_a(s) = \frac{E_a(s) - E_b(s)}{(R_a + sL_a)}$$

Since the field current is kept constant, the torque developed is proportional to the armature current, *i.e.*,

$$T_m(t) = K_t\, i_a(t)$$

Taking Laplace Transform of equation (5.28), we have,

$$T_m(s) = K_t\, I_a(s)$$

The torque equation is given by

$$T_m(t) = J\frac{d^2\theta(t)}{dt^2} + f\frac{d\theta(t)}{dt}$$

Taking Laplace Transform of equation

$$T_m(s) = Js^2\theta(s) + fs\theta(s)$$

$$\theta(s)(sf + s^2 J) = K_t I_a(s)$$

$$s\theta(s)[f + sJ] = K_t\frac{[E_a(s) - E_b(s)]}{(R_a + sL_a)}$$

Motor back e.m.f. is proportional to speed, *i.e.*,

$$e_b(t) = K_b\,.\frac{d\theta}{dt}$$

where K_b = back emf = constant.

Hence $$E_b(s) = sK_b\,.\,\theta(s)$$

Therefore, $$s\theta(s)(f + sJ) = \frac{K_t[E_a(s) - sK_b\,.\,\theta(s)]}{(R_a + sL_a)}$$

$$s\theta(s)(f + sJ)(R_a + sL_a) + sK_t\,.\,K_b\,.\,\theta(s) = K_t\,.\,E_a(s)$$

Transfer function is given by

$$G(s) = \frac{\theta(s)}{E_a(s)}$$

$$= \frac{K_t}{sK_t K_b + s(f + sJ)(R_a + sL_a)}$$

$$G(s) = \frac{K_t}{s[K_t K_b + f(1 + s\tau_{me})R_a(1 + s\tau_a)]}$$

where $$\tau_{me} = \text{Mechanical time constant} = \frac{J}{f}$$

$$\tau_a = \text{armature time constant} = \frac{L_a}{R_a}$$

Hence $$G(s) = \frac{K_t}{s[K_t K_b + R_a f(1 + s\tau_{me})(1 + s\tau_a)]}$$

Usually for an armature controlled machine, $\tau_a = 0$.

Hence we can have,

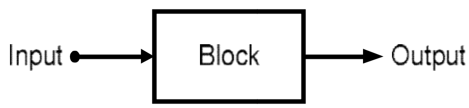
$$G(s) = \frac{K_t}{s[K_t K_b + fR_a(1 + s\tau_{me})]}$$

CHAPTER-04

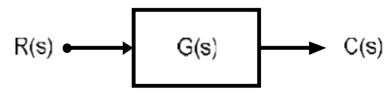
BLOCK DIAGRAM ALGEBRA & SIGNAL FLOW GRAPHS

BASIC DEFINITION IN BLOCK DIAGRAM MODEL:

Block diagram: It is the pictorial representation of the cause-and-response relationship between input and output of a physical system.



(a)



(b)

Fig. (a) A block diagram representation of a system and
(b) A block diagram representation with gain of a system

Output: The value of input multiplied by the gain of the system.

$$C(s) = G(s)R(s)$$

Summing point: It is the component of a block diagram model at which two or more signals can be added or subtracted. In Fig, inputs R(s) and B(s) have been given to a summing point and its output signal is E(s). Here,

$$E(s) = R(s) - B(s)$$

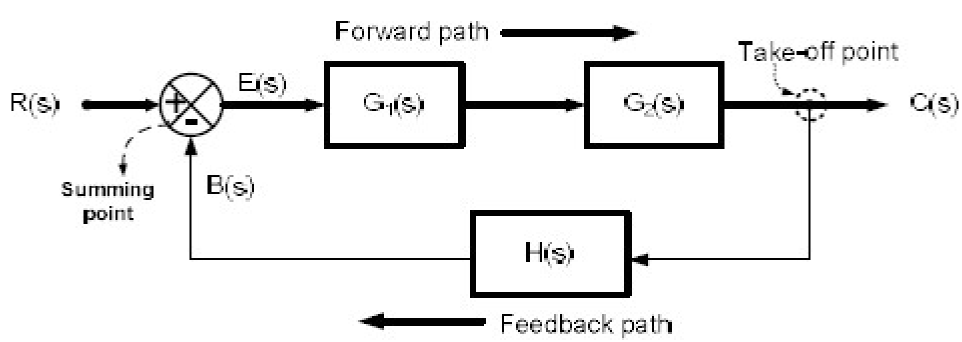


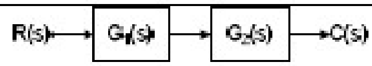
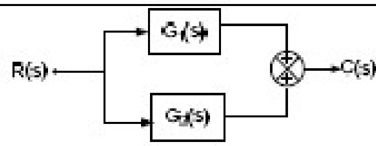
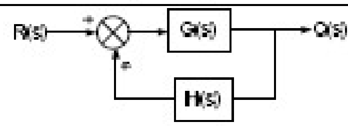
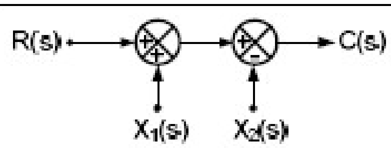
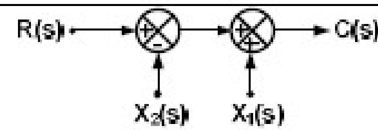
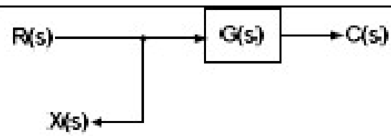
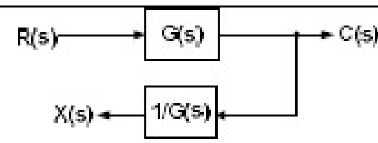
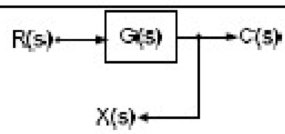
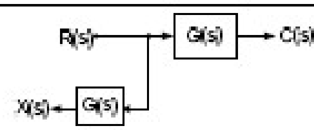
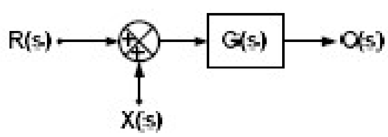
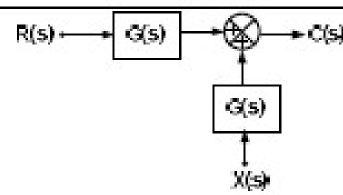
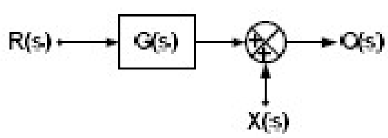
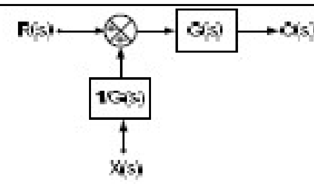
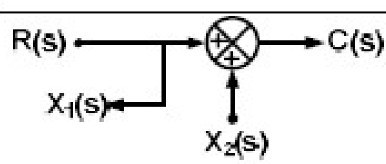
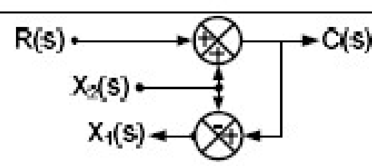
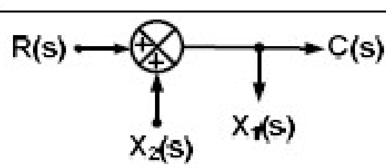
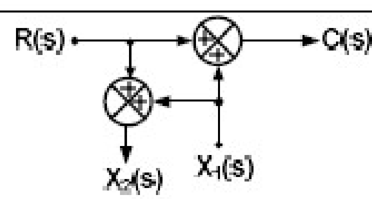
Fig. A block diagram representation of a system showing its different components

Take-off point: It is the component of a block diagram model at which a signal can be taken directly and supplied to one or more points as shown in Fig.

Forward path: It is the direction of signal flow from input towards output.

Feedback path: It is the direction of signal flow from output towards input.

RULES FOR REDUCTION OF BLOCK DIAGRAM MODEL:

Sl. No.	Rule No.	Configuration	Equivalent	Name
1	Rule 1			Cascade
2	Rule 2			Parallel
3	Rule 3			Loop
4	Rule 4			Associative Law
5	Rule 5			Move take-off point after a block
6	Rule 6			Move take-off point before a block
7	Rule 7			Move summing-point point after a block
8	Rule 8			Move summing-point point before a block
9	Rule 9			Move take-off point after a summing-point
10	Rule 10			Move take-off point before a summing-point

PROCEDURE FOR REDUCTION OF BLOCK DIAGRAM MODEL:

Step 1: Reduce the cascade blocks.

Step 2: Reduce the parallel blocks.

Step 3: Reduce the internal feedback loops.

Step 4: Shift take-off points towards right and summing points towards left.

Step 5: Repeat step 1 to step 4 until the simple form is obtained.

Step 6: Find transfer function of whole system as $\frac{C(s)}{R(s)}$.

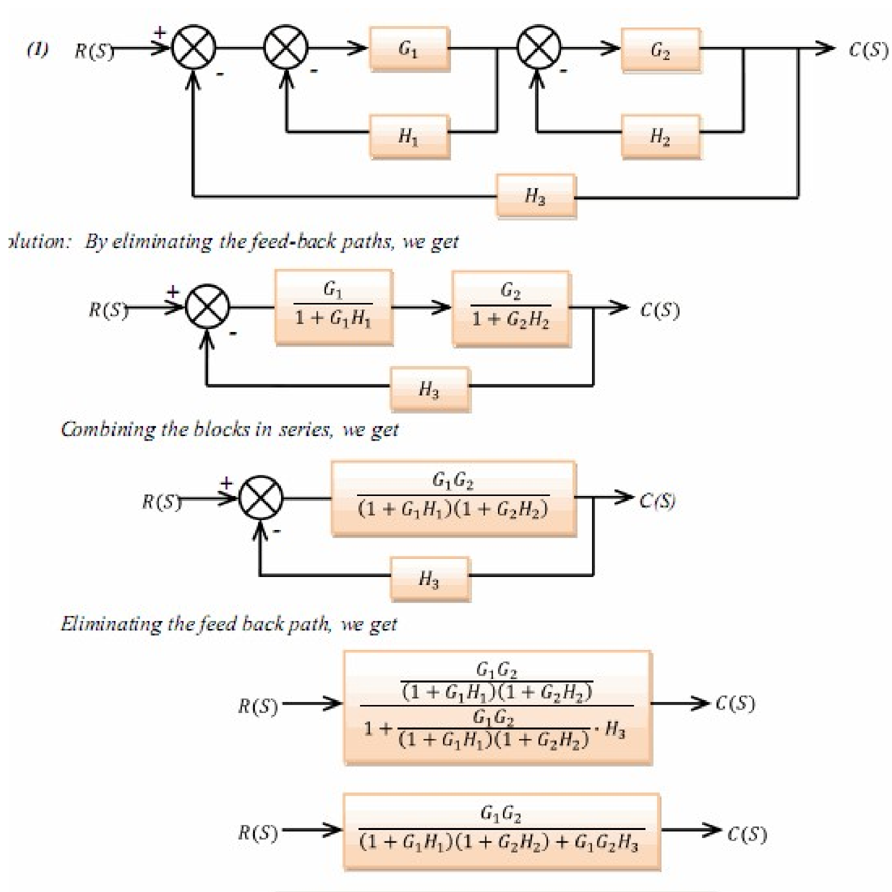
PROCEDURE FOR FINDING OUTPUT OF BLOCK DIAGRAM MODEL WITH MULTIPLE INPUTS:

Step 1: Consider one input taking rest of the inputs zero, find output.

Step 2: Follow step 1 for each inputs of the given Block Diagram model and find their corresponding outputs.

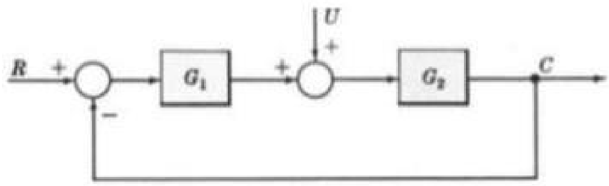
Step 3: Find the resultant output by adding all individual outputs.

Example 1:- Find C(s)/R(s) using block diagram reduction rules

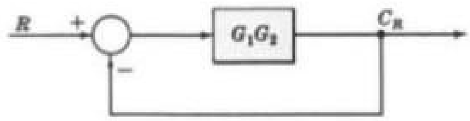


$$\Rightarrow \boxed{TF = \frac{C(S)}{R(S)} = \frac{G_1G_2}{(1+G_1H_1)(1+G_2H_2)+G_1G_2H_3}}$$

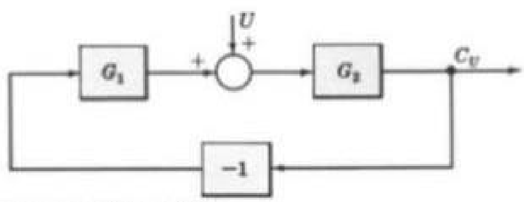
Example-2:- Determine output C due to inputs R & U using the superposition method.



Step 1: Put $U \equiv 0$.
Step 2: The system reduces to

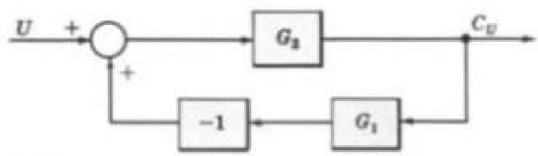


Step 3: the output C_R due to input R is $C_R = [G_1G_2/(1 + G_1G_2)]R$.

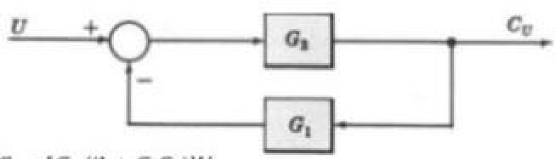


Step 4a: Put $R = 0$.
Step 4b: Put -1 into a block, representing the negative feedback effect:

Rearrange the block diagram:



Let the -1 block be absorbed into the summing point:



Step 4c: the output C_U due to input U is $C_U = [G_2/(1 + G_1G_2)]U$.

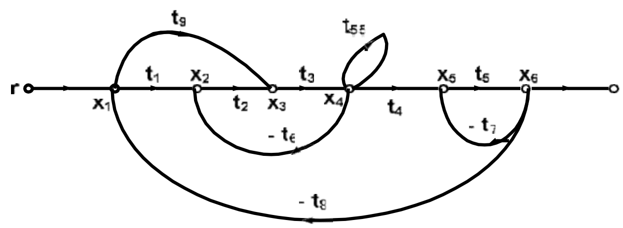
Step 5: The total output is $C = C_R + C_U$

$$= \left[\frac{G_1G_2}{1 + G_1G_2}\right]R + \left[\frac{G_2}{1 + G_1G_2}\right]U$$

$$= \left[\frac{G_2}{1 + G_1G_2}\right][G_1R + U]$$

SIGNAL FLOW GRAPHS (SFGS)

It is a pictorial representation of a system that graphically displays the signal transmission in it.



Basic Definitions in SFGs:

Input or source node: It is a node that has only outgoing branches i.e. node 'r'.

Output or sink node: It is a node that has only incoming branches i.e. node 'c'.

Chain node: It is a node that has both incoming and outgoing branches i.e. nodes 'x_1', 'x_2','x_3','x_4','x_5'and 'x_6'.

Gain or transmittance: It is the relationship between variables denoted by two nodes or value of a branch., Transmittances are 't_1', 't_2','t_3','t_4','t_5'and 't_6'.

Forward path: It is a path from input node to output node without repeating any of the nodes in between them. There are two forward paths, i.e. path-1:'r-x_1-x_2-x_3-x_4-x_5-x_6-c' and path-2:'r-x_1-x_3-x_4-x_5-x_6-c'.

Feedback path: It is a path from output node or a node near output node to a node near input node without repeating any of the nodes in between them.

Loop: It is a closed path that starts from one node and reaches the same node after trading through other nodes. There are four loops, i.e. loop-1:'x_2-x_3-x_4-x_1', loop-2:'x_5-x_6- x_5', loop-3:'x_1-x_2-x_3-x_4-x_5-x_6-x_1' and loop-4:'x_1-x_3-x_4-x_5-x_6-x_1'.

Self Loop: It is a loop that starts from one node and reaches the same node without trading through other nodes i.e. loop in node 'x_4' with transmittance 't_{55}'.

Path gain: It is the product of gains or transmittances of all branches of a forward path. The path gains are $P_1 = t_1t_2t_3t_4t_5$ (for path-1) and $P_2 = t_9t_3t_4t_5$ (for path-2).

Loop gain: It is the product of gains or transmittances of all branches of a loop. There are four loops, i.e. $L_1 = -t_2t_3t_6$, $L_2 = -t_5t_7$, $L_3 = -t_1t_2t_3t_4t_5t_8$, and $L_4 = -t_9t_3t_4t_5t_8$.

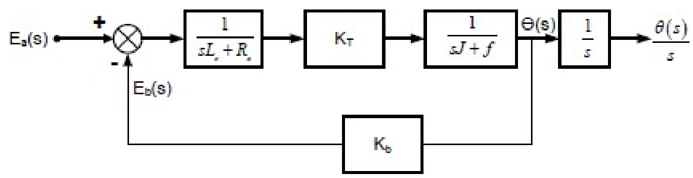
Dummy node: If the first node is not an input node and/or the last node is not an output node than a node is connected before the existing first node and a node is connected after the existing last node with unity transmittances. These nodes are called dummy nodes. 'r' and 'c' are the dummy nodes.

Non-touching Loops: Two or more loops are non-touching loops if they don't have any common nodes between them. L_1 and L_2 are non-touching loops

PROPERTIES OF SFGS:

- Applied to linear system
- Arrow indicates signal flow
- Nodes represent variables, summing points and take-off points
- Algebraic sum of all incoming signals and outgoing nodes is zero
- SFG of a system is not unique
- Overall gain of an SFG can be determined by using Mason's gain formula

SFG FROM BLOCK DIAGRAM MODEL:

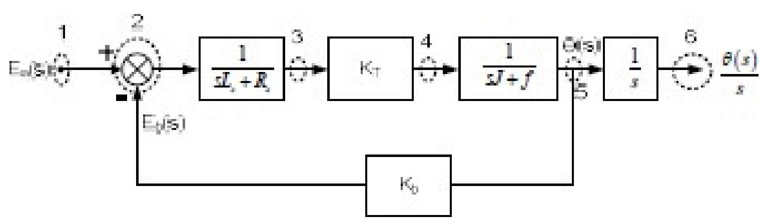


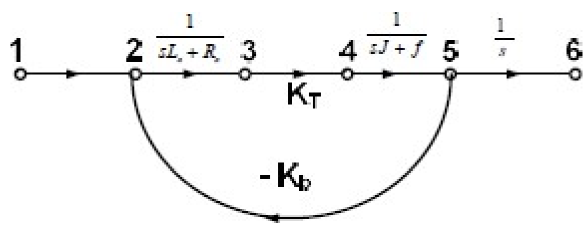
Step-1: All variables and signals are replaced by nodes.

Step-2: Connect all nodes according to their signal flow.

Step-3: Each of gains is replaced by transmittances of the branches connected between two nodes of the forward paths.

Step-4: Each of gains is replaced by transmittances multiplied with (-1) of the branches connected between two nodes of the forward paths.





MASON'S GAIN FORMULA:

Transfer function of a system=

$$G(s) = \frac{C(s)}{R(s)} = \frac{\sum_{k=1}^{N} P_k \Delta_k}{\Delta}$$

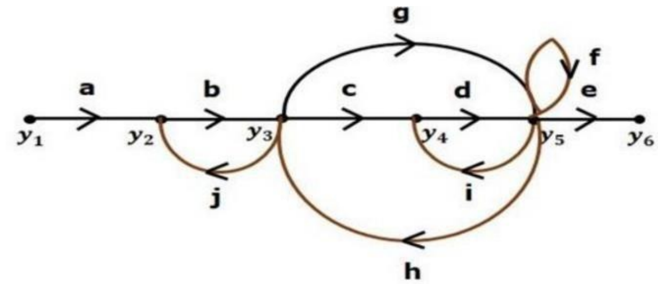
N= total number of forwardpaths

P_k= path gain of k^{th} forward path

Δ= 1 - ($\sum$loop gains of all individual loops) + ($\sum$gain product of loop gains of all possible two non-touching loops) - ($\sum$gain product of loop gains of all possible three non-touching loops) + …

Δ_k= value of Δ after eliminating all loops that touches k^{th} forward path

Example:- Find overall transfer function of system using Mason's gain formula



- Number of forward paths, N = 2.
- First forward path is - $y_1 \rightarrow y_2 \rightarrow y_3 \rightarrow y_4 \rightarrow y_5 \rightarrow y_6$.
- First forward path gain, $p_1 = abcde$.
- Second forward path is - $y_1 \rightarrow y_2 \rightarrow y_3 \rightarrow y_5 \rightarrow y_6$.
- Second forward path gain, $p_2 = abge$.
- Number of individual loops, L = 5.
- Loops are - $y_2 \rightarrow y_3 \rightarrow y_2$, $y_3 \rightarrow y_5 \rightarrow y_3$, $y_3 \rightarrow y_4 \rightarrow y_5 \rightarrow y_3$,

  $y_4 \rightarrow y_5 \rightarrow y_4$ and $y_5 \rightarrow y_5$.
- Loop gains are - $l_1 = bj$, $l_2 = gh$, $l_3 = cdh$, $l_4 = di$ and $l_5 = f$.
- Number of two non-touching loops = 2.
- First non-touching loops pair is - $y_2 \rightarrow y_3 \rightarrow y_2$, $y_4 \rightarrow y_5 \rightarrow y_4$.
- Gain product of first non-touching loops pair, $l_1 l_4 = bjdi$
- Second non-touching loops pair is - $y_2 \rightarrow y_3 \rightarrow y_2$, $y_5 \rightarrow y_5$.
- Gain product of second non-touching loops pair is - $l_1 l_5 = bjf$

Higher number of (more than two) non-touching loops are not present in this signal flow graph.

We know,

$$\Delta = 1 - (sum\ of\ all\ individual\ loop\ gains)$$
$$+(sum\ of\ gain\ products\ of\ all\ possible\ two\ nontouching\ loops)$$
$$-(sum\ of\ gain\ products\ of\ all\ possible\ three\ nontouching\ loops) + \ldots$$

Substitute the values in the above equation,

$$\Delta = 1 - (bj + gh + cdh + di + f) + (bjdi + bjf) - (0)$$

$$\Rightarrow \Delta = 1 - (bj + gh + cdh + di + f) + bjdi + bjf$$

There is no loop which is non-touching to the first forward path.

So, $\Delta_1 = 1$.

Similarly, $\Delta_2 = 1$. Since, no loop which is non-touching to the second forward path.

Substitute, N = 2 in Mason's gain formula

$$T = \frac{C(s)}{R(s)} = \frac{\Sigma_{i=1}^{2} P_i \Delta_i}{\Delta}$$

$$T = \frac{C(s)}{R(s)} = \frac{P_1 \Delta_1 + P_2 \Delta_2}{\Delta}$$

Substitute all the necessary values in the above equation.

$$T = \frac{C(s)}{R(s)} = \frac{(abcde)1 + (abge)1}{1 - (bj + gh + cdh + di + f) + bjdi + bjf}$$

$$\Rightarrow T = \frac{C(s)}{R(s)} = \frac{(abcde) + (abge)}{1 - (bj + gh + cdh + di + f) + bjdi + bjf}$$

Therefore, the transfer function is -

$$T = \frac{C(s)}{R(s)} = \frac{(abcde) + (abge)}{1 - (bj + gh + cdh + di + f) + bjdi + bjf}$$

CONSTRUCTION OF SIGNAL FLOW GRAPH FROM ALGEBRAIC EQUATIONS:-

Let us construct a signal flow graph by considering the following algebraic equations

$$y_2 = a_{12}y_1 + a_{42}y_4$$

$$y_3 = a_{23}y_2 + a_{53}y_5$$

$$y_4 = a_{34}y_3$$

$$y_5 = a_{45}y_4 + a_{35}y_3$$

$$y_6 = a_{56}y_5$$

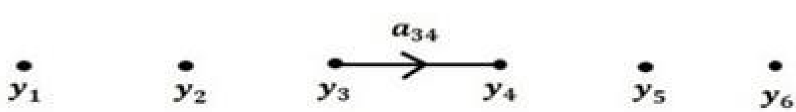
There will be six **nodes** (y_1, y_2, y_3, y_4, y_5 and y_6) and eight **branches** in this signal flow graph. The gains of the branches are a_{12}, a_{23}, a_{34}, a_{45}, a_{56}, a_{42}, a_{53} and a_{35}.

To get the overall signal flow graph, draw the signal flow graph for each equation, then combine all these signal flow graphs and then follow the steps given below −

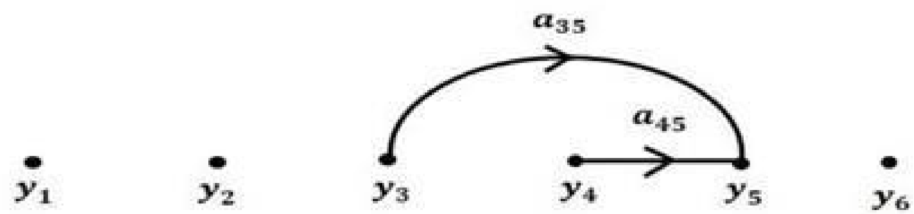
Step 1 − Signal flow graph for $y_2 = a_{13}y_1 + a_{42}y_4$ is shown in the following figure.

Step 2 − Signal flow graph for $y_3 = a_{23}y_2 + a_{53}y_5$ is shown in the following figure.

Step 3 − Signal flow graph for $y_4 = a_{34}y_3$ is shown in the following figure.

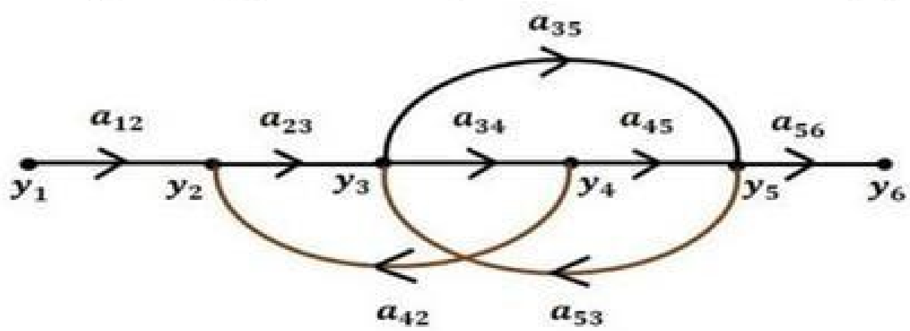


Step 4 – Signal flow graph for $y_5 = a_{45}y_4 + a_{35}y_3$ is shown in the following figure.



Step 5 – Signal flow graph for $y_6 = a_{56}y_5$ is shown in the following figure.

Step 6 – Signal flow graph of overall system is shown in the following figure.

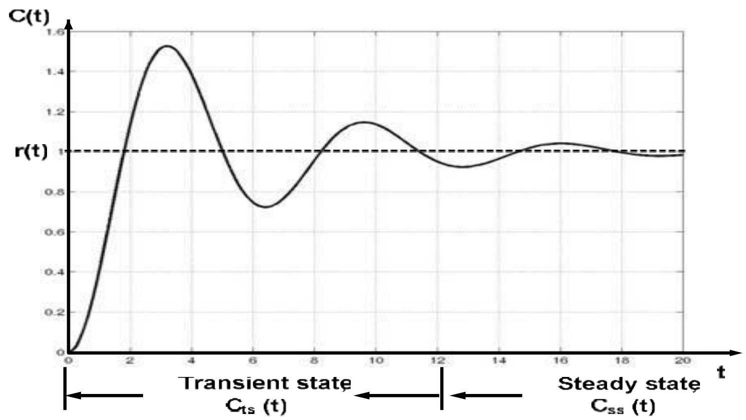


CHAPTER-05

TIME RESPONSE ANALYSIS

TIME RESPONSE OF CONTROL SYSTEM:

Time response c(t)is the variation of output with respect to time. The part of time response that goes to zero after large interval of time is called transient response $c_{tr}(t)$. The part of time response that remains after transient response is called steady-state response $c_{ss}(t)$.



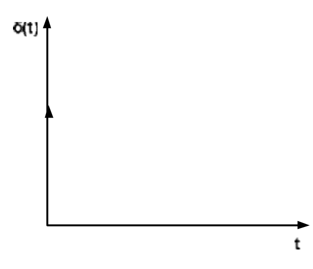
STANDARD TEST SIGNALS

1. Impulse Signal: An impulse signal δ(t) is mathematically defined as follows.

$$\delta(t) = \left.\begin{matrix} undefined & ;t=0 \\ 0 & ;t\neq 0 \end{matrix}\right\}$$

Laplace transform of impulse signal is

$$\delta(s) = 1$$

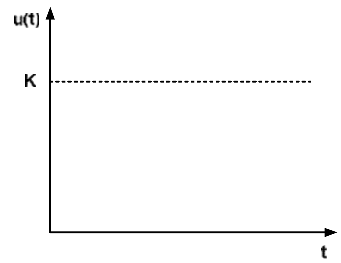


2. Step Signal: A step signal u(t) is mathematically defined as follows.

$$u(t) = \left.\begin{matrix} 0 & ;t<0 \\ K & ;t\geq 0 \end{matrix}\right\}$$

Laplace transform of step signal is

$$U(s) = \frac{K}{s}$$



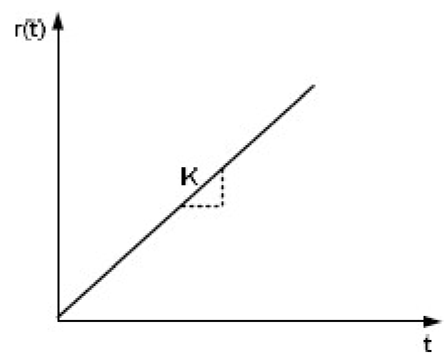
3. Ramp Signal: A step signal r(t) is mathematically defined as follows.

$$r(t) = \begin{matrix} 0 & ;t<0 \\ Kt & ;t \geq 0 \end{matrix} \Bigg\} \quad (8.10)$$

Laplace transform of ramp signal

is

$$R(s) = \frac{K}{s^2}$$

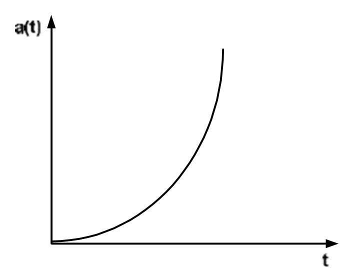


4. Parabolic Signal A step signal a(t) is mathematically defined as follows.

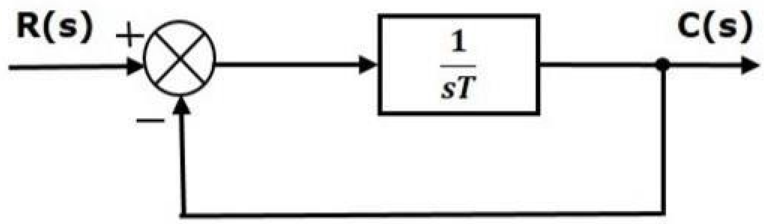
$$a(t) = \begin{matrix} 0 & ;t<0 \\ \frac{Kt^2}{2} & ;t \geq 0 \end{matrix} \Bigg\}$$

Laplace transform of parabolic signal is

$$A(s) = \frac{K}{s^3}$$



TIME RESPONSE OF 1ST ORDER SYSTEM:



We know that the transfer function of the closed loop control system has unity negative feedback as,

$$\frac{C(s)}{R(s)} = \frac{G(s)}{1+G(s)}$$

Substitute, $G(s) = \frac{1}{sT}$ in the above equation.

$$\frac{C(s)}{R(s)} = \frac{\frac{1}{sT}}{1+\frac{1}{sT}} = \frac{1}{sT+1}$$

The power of s is one in the denominator term. Hence, the above transfer function is of the first order and the system is said to be the **first order system**.

We can re-write the above equation as

$$C(s) = \left(\frac{1}{sT+1}\right) R(s)$$

Where,

- **C(s)** is the Laplace transform of the output signal c(t),
- **R(s)** is the Laplace transform of the input signal r(t), and
- **T** is the time constant.

Follow these steps to get the response (output) of the first order system in the time domain.

- Take the Laplace transform of the input signal $r(t)$.
- Consider the equation, $C(s) = \left(\frac{1}{sT+1}\right) R(s)$

- Substitute $R(s)$ value in the above equation.
- Do partial fractions of $C(s)$ if required.
- Apply inverse Laplace transform to $C(s)$.

(i) Unit Step Response:-

Consider the **unit step signal** as an input to first order system.

So, r(t)=u(t)

$$R(s) = \frac{1}{s}$$

Consider the equation, $C(s) = \left(\frac{1}{sT+1}\right) R(s)$

Substitute, $R(s) = \frac{1}{s}$ in the above equation.

$$C(s) = \left(\frac{1}{sT+1}\right)\left(\frac{1}{s}\right) = \frac{1}{s\,(sT+1)}$$

Do partial fractions of C(s).

$$C(s) = \frac{1}{s\,(sT+1)} = \frac{A}{s} + \frac{B}{sT+1}$$

$$\Rightarrow \frac{1}{s\,(sT+1)} = \frac{A\,(sT+1)+Bs}{s\,(sT+1)}$$

On both the sides, the denominator term is the same. So, they will get cancelled by each other. Hence, equate the numerator terms.

$$1=A(sT+1)+Bs$$

By equating the constant terms on both the sides, you will get A = 1.
Substitute, A = 1 and equate the coefficient of the **s** terms on both the sides.

$$0=T+B$$
$$\Rightarrow B=-T$$

Substitute, A = 1 and B = −T in partial fraction expansion of C(s)

$$C(s) = \frac{1}{s} - \frac{T}{sT+1} = \frac{1}{s} - \frac{T}{T\left(s+\frac{1}{T}\right)}$$

$$\Rightarrow C(s) = \frac{1}{s} - \frac{1}{s+\frac{1}{T}}$$

Apply inverse Laplace transform on both the sides.

$$c(t) = \left(1 - e^{-\left(\frac{t}{T}\right)}\right) u(t)$$

The **unit step response**, c(t) has both the transient and the steady state terms.

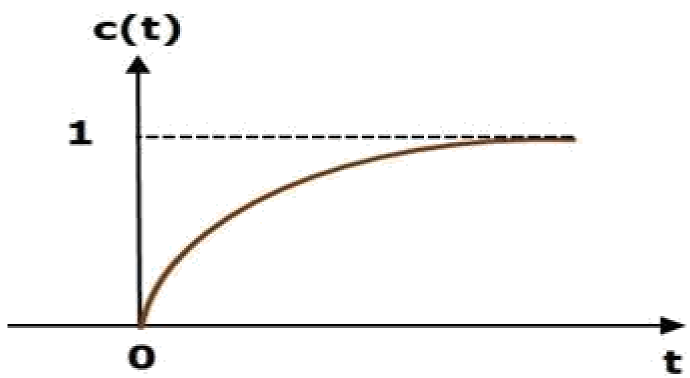
The transient term in the unit step response is -

$$c_{tr}(t) = -e^{-\left(\frac{t}{T}\right)} u(t)$$

The steady state term in the unit step response is –

$$c_{ss}(t) = u(t)$$

The following figure shows the unit step response



The value of the **unit step response, c(t)** is zero at t = 0 and for all negative values of t. It is gradually increasing from zero value and finally reaches to one in steady state. So, the steady state value depends on the magnitude of the input.

(ii) Unit impulse response:

Consider the **unit impulse signal** as an input to the first order system.

So, $r(t)=\delta(t)$

Apply Laplace transform on both the sides.

$R(s) = 1$

Consider the equation, $C(s) = \left(\frac{1}{sT+1}\right) R(s)$

Substitute, $R(s) = 1$ in the above equation.

$$C(s) = \left(\frac{1}{sT+1}\right)(1) = \frac{1}{sT+1}$$

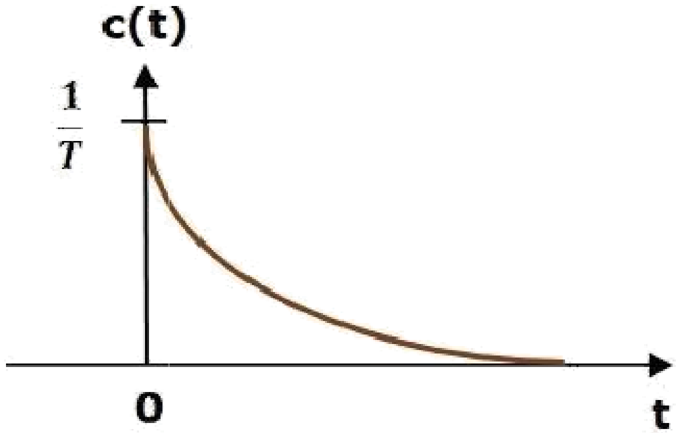
Rearrange the above equation in one of the standard forms of Laplace transforms.

$$C(s) = \frac{1}{T\left(s+\frac{1}{T}\right)} \Rightarrow C(s) = \frac{1}{T}\left(\frac{1}{s+\frac{1}{T}}\right)$$

Applying Inverse Laplace Transform on both the sides,

$$c(t) = \frac{1}{T} e^{\left(-\frac{t}{T}\right)} u(t)$$

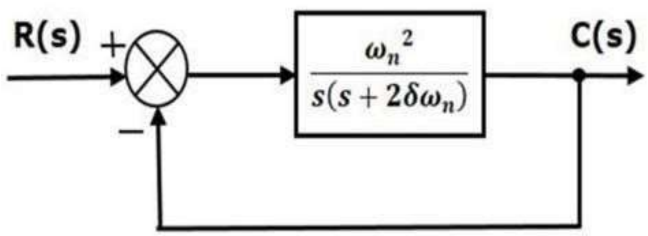
The unit impulse response is shown in the following figure.



The **unit impulse response**, c(t) is an exponential decaying signal for positive values of 't' and it is zero for negative values of 't'.

TIME RESPONSE OF 2ND ORDER SYSTEM

Consider the following block diagram of closed loop control system. Here, an open loop transfer function, ωn^2 / s(s+2δωn) is connected with a unity negative feedback.



We know that the transfer function of the closed loop control system having unity negative feedback as

$$\frac{C(s)}{R(s)} = \frac{G(s)}{1+G(s)}$$

Substitute, $G(s) = \frac{\omega_n^2}{s(s+2\delta\omega_n)}$ in the above equation.

$$\frac{C(s)}{R(s)} = \frac{\left(\frac{\omega_n^2}{s(s+2\delta\omega_n)}\right)}{1+\left(\frac{\omega_n^2}{s(s+2\delta\omega_n)}\right)} = \frac{\omega_n^2}{s^2+2\delta\omega_n s+\omega_n^2}$$

STEP RESPONSE OF 2^{ND} ORDER SYSTEM

Consider the unit step signal as an input to the second order system. Laplace transform of the unit step signal is,

$$T(s)=\frac{C(s)}{R(s)}=\frac{\omega_n^2}{s^2+2\zeta\omega_n s+\omega_n^2}$$

$$Now,\ r(t)=1\ or\ R(s)=\frac{1}{s}$$

$$\therefore C(s)=\frac{1}{s}\cdot\frac{\omega_n^2}{s^2+2\zeta\omega_n s+\omega_n^2}$$

$$=\frac{1}{s}\cdot\frac{\omega_n^2}{s^2+2\zeta\omega_n s+\zeta^2\omega_n^2+\omega_n^2-\zeta^2\omega_n^2}$$

$$=\frac{1}{s}\cdot\frac{\omega_n^2}{(s+\zeta\omega_n)^2+\omega_n^2(1-\zeta^2)}$$

$$=\frac{(s+\zeta\omega_n)^2+\omega_n^2(1-\zeta^2)-(s+\zeta\omega_n)^2+\omega_n^2\zeta^2}{s\left\{(s+\zeta\omega_n)^2+\omega_n^2(1-\zeta^2)\right\}}$$

$$=\frac{(s+\zeta\omega_n)^2+\omega_n^2(1-\zeta^2)-s^2-2s\zeta\omega_n-\omega_n^2\zeta^2+\omega_n^2\zeta^2}{s\left\{(s+\zeta\omega_n)^2+\omega_n^2(1-\zeta^2)\right\}}$$

$$=\frac{(s+\zeta\omega_n)^2+\omega_n^2(1-\zeta^2)-s(s+2s\zeta\omega_n)}{s\left\{(s+\zeta\omega_n)^2+\omega_n^2(1-\zeta^2)\right\}}$$

$$=\frac{1}{s}-\frac{s+2s\zeta\omega_n}{(s+\zeta\omega_n)^2+\omega_n^2(1-\zeta^2)}$$

$$Putting,\ \omega_d=\omega_n\sqrt{1-\zeta^2}$$

$$=\frac{1}{s}-\frac{s+\zeta\omega_n}{(s+\zeta\omega_n)^2+\omega_d^2}-\frac{\zeta\omega_n}{(s+\zeta\omega_n)^2+\omega_d^2}$$

$$=\frac{1}{s}-\frac{s+\zeta\omega_n}{(s+\zeta\omega_n)^2+\omega_d^2}-\frac{\zeta\omega_n}{\omega_d}\cdot\frac{\omega_d}{(s+\zeta\omega_n)^2+\omega_d^2}$$

Taking the inverse Laplace Transform of above equation, we get

$$\pounds^{-1}\left[C(s)\right]=\pounds^{-1}\left[\frac{1}{s}-\frac{s+\zeta\omega_n}{(s+\zeta\omega_n)^2+\omega_d^2}-\frac{\zeta\omega_n}{\omega_d}\cdot\frac{\omega_d}{(s+\zeta\omega_n)^2+\omega_d^2}\right]$$

$$= \mathcal{L}^{-1}\left[\frac{1}{s}\right] - \mathcal{L}^{-1}\left[\frac{s+\zeta\omega_n}{(s+\zeta\omega_n)^2+\omega_d^2}\right] - \mathcal{L}^{-1}\left[\frac{\zeta\omega_n}{\omega_d}\cdot\frac{\omega_d}{(s+\zeta\omega_n)^2+\omega_d^2}\right]$$

$$\therefore c(t) = 1 - e^{-\zeta\omega_n t}\cdot\cos\omega_d t - \frac{\zeta\omega_n}{\omega_d}\cdot e^{-\zeta\omega_n t}\cdot\sin\omega_d t$$

$$\because \mathcal{L}^{-1}\left[\frac{1}{s}\right] = 1,\ \mathcal{L}^{-1}\left[\frac{s+\alpha}{(s+\alpha)^2+\omega^2}\right] = e^{-\alpha t}\cos\omega t,$$

$$\mathcal{L}^{-1}\left[\frac{\omega}{(s+\alpha)^2+\omega^2}\right] = e^{-\alpha t}\sin\omega t$$

The above expression of output c(t) can be rewritten as

$$c(t) = 1 - e^{-\zeta\omega_n t}\left(\cos\omega_d t + \frac{\zeta}{\sqrt{1-\zeta^2}}\cdot\sin\omega_d t\right)$$

$$= 1 - \frac{e^{-\zeta\omega_n t}}{\sqrt{1-\zeta^2}}\left(\sqrt{1-\zeta^2}\cos\omega_d t + \zeta\cdot\sin\omega_d t\right)$$

$$\left[Say,\ \zeta = \cos\phi,\ hence,\ \sqrt{1-\zeta^2} = \sin\phi\right]$$

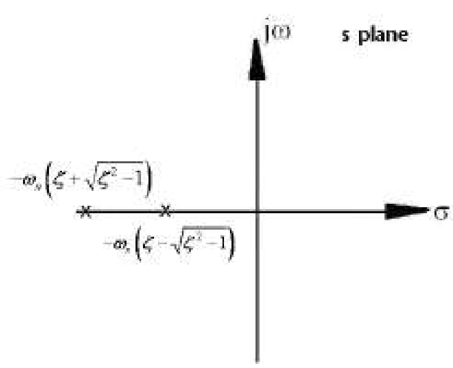
$$\therefore c(t) = 1 - \frac{e^{-\zeta\omega_n t}}{\sqrt{1-\zeta^2}}\left(\sin\phi\cos\omega_d t + \cos\phi\sin\omega_d t\right)$$

$$= 1 - \frac{e^{-\zeta\omega_n t}}{\sqrt{1-\zeta^2}}\sin\left(\omega_d t + \phi\right)$$

(a) ζ>1 over damped
Poles are:

$$s_{1,2} = -\omega_n\left(\zeta \pm \sqrt{\zeta^2 - 1}\right)$$

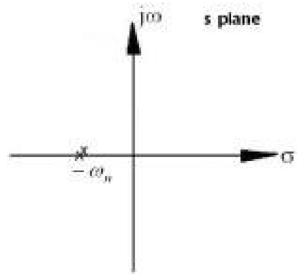
Graphically, the poles of an over damped system is shown as follows.



(b) ζ =1 critically damped
Poles are:

$$s_{1,2} = -\omega_n$$

Graphically, the poles of an critically damped system is shown as follows.



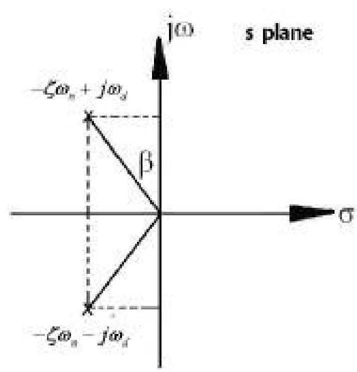
(c) ζ<1 under damped
Poles are:

$$s_{1,2} = -\omega_n\left(\zeta \pm \mathrm{j}\sqrt{1-\zeta^2}\right)$$
$$\Rightarrow s_{1,2} = -\zeta\omega_n \pm j\omega_d$$

Where, ω_d = Damped natural frequency

$$\omega_d = \omega_n\sqrt{1-\zeta^2}$$

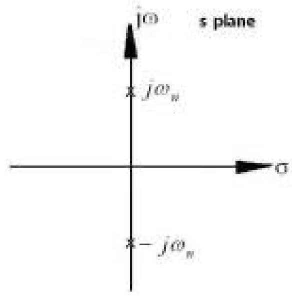
Graphically, the poles of an critically damped system is shown as follows.



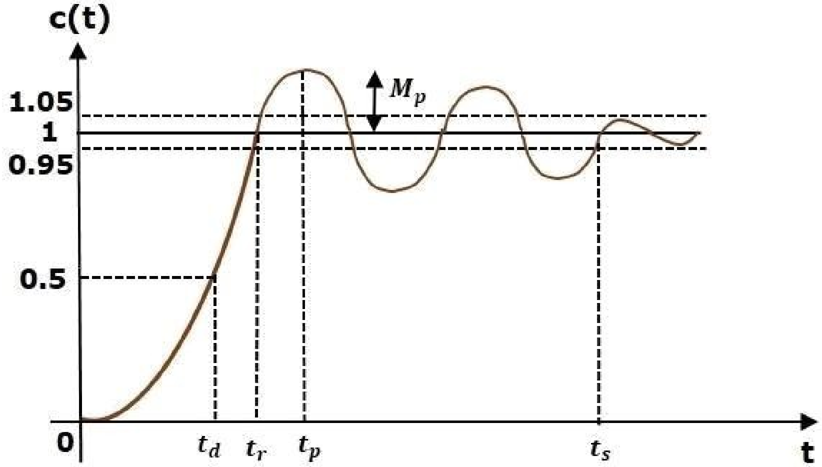
Here, $\tan\beta = \dfrac{\zeta}{\sqrt{1-\zeta^2}}$

(d) $\zeta = 0$ un-damped
Poles are:

$$s_{1,2} = -\pm\,\mathrm{j}\,\omega_n$$



TIME RESPONSE SPECIFICATION



Delay Time

It is the time required for the response to reach **half of its final value** from the zero instant. It is denoted by t_d.

Rise Time

It is the time required for the response to rise from **0% to 100% of its final value**. This is applicable for the **under-damped systems**. For the over-damped systems, consider the duration from 10% to 90% of the final value. Rise time is denoted by **tr**.

As per definition, the magnitude of output signal at Rise times is 1. That is c(t) = 1, hence

$$1 = 1 - \frac{e^{-\zeta\omega_n t_r}}{\sqrt{1-\zeta^2}} \sin\left\{\left(\omega_n\sqrt{1-\zeta^2}\right)t_r + \phi\right\}$$

$$\Rightarrow \frac{e^{-\zeta\omega_n t_r}}{\sqrt{1-\zeta^2}} \sin\left\{\left(\omega_n\sqrt{1-\zeta^2}\right)t_r + \phi\right\} = 0$$

$$\Rightarrow \sin\left\{\left(\omega_n\sqrt{1-\zeta^2}\right)t_r + \phi\right\} = 0$$

$$\Rightarrow \left\{\left(\omega_n\sqrt{1-\zeta^2}\right)t_r + \phi\right\} = \pi$$

$$\Rightarrow t_r = \frac{\pi - \phi}{\omega_n\sqrt{1-\zeta^2}}$$

Peak Time

It is the time required for the response to reach the **peak value** for the first time. It is denoted by tp. At t=tp the first derivate of the response is zero.

As per definition at the peak time, the response curve reaches to its maximum value. Hence at that point,

$$\frac{\mathrm{d}c(t)}{\mathrm{d}t} = 0$$

$$Now,\ c(t) = 1 - \frac{e^{-\zeta\omega_n t}}{\sqrt{1-\zeta^2}} \sin\left\{\left(\omega_n\sqrt{1-\zeta^2}\right)t + \phi\right\}$$

$$\therefore \frac{\mathrm{d}c(t)}{\mathrm{d}t} = -\frac{e^{-\zeta\omega_n t}}{\sqrt{1-\zeta^2}} \cdot \omega_n\sqrt{1-\zeta^2}\cos\left\{\left(\omega_n\sqrt{1-\zeta^2}\right)t + \phi\right\}$$

$$-\frac{(-\zeta\omega_n)\,e^{-\zeta\omega_n t}}{\sqrt{1-\zeta^2}} \sin\left\{\left(\omega_n\sqrt{1-\zeta^2}\right)t + \phi\right\}$$

$$Putting,\ \frac{\mathrm{d}c(t)}{\mathrm{d}t} = 0$$

$$\therefore \frac{e^{-\zeta\omega_n t}}{\sqrt{1-\zeta^2}}\left[-\omega_n\sqrt{1-\zeta^2}\cos\left\{\left(\omega_n\sqrt{1-\zeta^2}\right)t + \phi\right\} + \zeta\omega_n\sin\left\{\left(\omega_n\sqrt{1-\zeta^2}\right)t + \phi\right\}\right]$$

$$= 0$$

$$\omega_n\sqrt{1-\zeta^2}\cos\left\{\left(\omega_n\sqrt{1-\zeta^2}\right)t + \phi\right\} = \zeta\omega_n\sin\left\{\left(\omega_n\sqrt{1-\zeta^2}\right)t + \phi\right\}$$

$$\Rightarrow \tan\left[\left(\omega_n\sqrt{1-\zeta^2}\right)t + \phi\right] = \frac{\sqrt{1-\zeta^2}}{\zeta} = \tan\phi$$

$$\therefore \left(\omega_n\sqrt{1-\zeta^2}\right)t = n\pi$$

$$Where,\ n = 1, 2, 3\ldots$$

$$t_p = \frac{\pi}{\omega_n\sqrt{1-\zeta^2}}$$

Peak Overshoot

Peak overshoot **Mp** is defined as the deviation of the response at peak time from the final value of response. It is also called the **maximum overshoot**.

Mathematically, we can write it as

$$M_p = c(t_p) - c(\infty)$$

Where, $c(t_p)$ is the peak value of the response, $c(\infty)$ is the final (steady state) value of the response.

$$c(t)_{max} = 1 - \frac{e^{-\zeta\omega_n t_p}}{\sqrt{1-\zeta^2}} \sin\left[\left(\omega_n\sqrt{1-\zeta^2}\right) t_p + \phi\right]$$

$$\Rightarrow c(t)_{max} = 1 - \frac{e^{-\zeta\omega_n \frac{\pi}{\omega_n\sqrt{1-\zeta^2}}}}{\sqrt{1-\zeta^2}} \sin\left[\left(\omega_n\sqrt{1-\zeta^2}\right)\frac{\pi}{\omega_n\sqrt{1-\zeta^2}} + \phi\right]$$

$$\Rightarrow c(t)_{max} = 1 - \frac{e^{\frac{-\zeta\pi}{\sqrt{1-\zeta^2}}}}{\sqrt{1-\zeta^2}} \sin(\pi+\phi) = 1 - \frac{e^{\frac{-\zeta\pi}{\sqrt{1-\zeta^2}}}}{\sqrt{1-\zeta^2}}(-\sin\phi)$$

$$= 1 + \frac{e^{\frac{-\zeta\pi}{\sqrt{1-\zeta^2}}}}{\sqrt{1-\zeta^2}}\sin\phi = 1 + \frac{e^{\frac{-\zeta\pi}{\sqrt{1-\zeta^2}}}}{\sqrt{1-\zeta^2}}\sqrt{1-\zeta^2} = 1 + e^{-\frac{\zeta\pi}{\sqrt{1-\zeta^2}}}$$

$$M_p = c(t)_{max} - 1 = \left(1 + e^{-\frac{\zeta\pi}{\sqrt{1-\zeta^2}}}\right) - 1$$

$$\Rightarrow M_p = e^{-\frac{\zeta\pi}{\sqrt{1-\zeta^2}}}$$

Settling time

It is the time required for the response to reach the steady state and stay within the specified tolerance bands around the final value. In general, the tolerance bands are 2% and 5%. The settling time is denoted by ts.

The settling time for 5% tolerance band is –

The settling time for 2% tolerance band is –

$$t_s = \frac{3}{\delta\omega_n} = 3\tau$$

$$t_s = \frac{4}{\delta\omega_n} = 4\tau$$

Where, τ is the time constant and is equal to 1/δωn.

- Both the settling time ts and the time constant τ are inversely proportional to the damping ratio δ.

Both the settling time ts and the time constant τ are independent of the system gain. That means even the system gain changes, the settling time ts and time constant τ will never change.

Steady state error:-

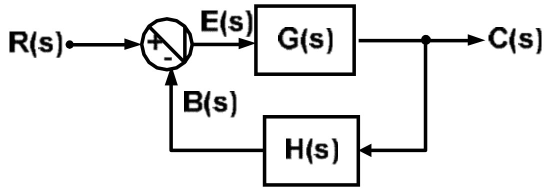
The deviation of the output of control system from desired response during steady state is known as **steady state error**. It is represented as e_{ss}. We can find steady state error using the final value theorem as follows.

$$e_{ss} = \lim_{t\to\infty} e(t) = \lim_{s\to 0} E(s)$$

Where,

E(s) is the Laplace transform of the error signal, e(t)

A simple closed-loop control system with negative feedback is shown as follows.



$E(s) = R(s) - B(s)$ --- (i)

$B(s) = C(s)\,H(s)$ ----(ii)

$C(s) = E(s)\,G(s)$---- (iii)

Applying value of B(s) of eq 2 into eq 1

$E(s) = R(s) - C(s)\,H(s)$

Applying value of C(s) of eq 3 into above eq

$$E(s) = R(s) - E(s)\,G(s)\,H(s)$$

$$\Rightarrow E(s) = \frac{R(s)}{1 + G(s)H(s)}$$

Steady-state error,

$$e_{ss} = \lim_{t \to \infty} e(t) = \lim_{s \to 0} sE(s)$$

$$e_{ss} = \lim_{s \to 0} sE(s) = \lim_{s \to 0} \frac{sR(s)}{1 + G(s)H(s)}$$

Therefore, steady-state error depends on two factors, i.e.

(a) type and magnitude of R(s)
(b) open-loop transfer function G(s)H(s)

Types of input and Steady-state error:

(i)Step Input:

$$R(s) = \frac{A}{s}$$

$$e_{ss} = \lim_{s\to 0}\frac{s\left(\frac{A}{s}\right)}{1+G(s)H(s)} = \lim_{s\to 0}\frac{A}{1+G(s)H(s)}$$

$$\Rightarrow e_{ss} = \frac{A}{1+\lim_{s\to 0}G(s)H(s)} = \frac{A}{1+K_P}$$

Where,

$$K_P = \lim_{s\to 0}G(s)H(s)$$

(ii) Ramp Input:

$$R(s) = \frac{A}{s^2}$$

$$e_{ss} = \lim_{s\to 0}\frac{s\left(\frac{A}{s^2}\right)}{1+G(s)H(s)} = \lim_{s\to 0}\frac{A}{s\left[1+G(s)H(s)\right]}$$

$$\Rightarrow e_{ss} = \lim_{s\to 0}\frac{A}{s+sG(s)H(s)}$$

$$\Rightarrow e_{ss} = \frac{A}{\lim_{s\to 0}sG(s)H(s)} = \frac{A}{K_V}$$

Where,

$$K_V = \lim_{s\to 0}sG(s)H(s)$$

(iii) Parabolic Input:

$$R(s) = \frac{A}{s^3}$$

$$e_{ss} = \lim_{s\to 0}\frac{s\left(\frac{A}{s^3}\right)}{1+G(s)H(s)} = \lim_{s\to 0}\frac{A}{s^2\left[1+G(s)H(s)\right]}$$

$$\Rightarrow e_{ss} = \lim_{s\to 0}\frac{A}{s^2+s^2G(s)H(s)}$$

$$\Rightarrow e_{ss} = \frac{A}{\lim_{s\to 0}s^2G(s)H(s)} = \frac{A}{K_A}$$

Where,

$$K_A = \lim_{s\to 0}s^2G(s)H(s)$$

Types of input and steady-state error are summarized as follows.

Error Constant	Equation	Steady-state error (e_{ss})
Position Error Constant (K_P)	$K_P = \lim_{s\to 0} G(s)H(s)$	$e_{ss} = \frac{A}{1+K_P}$
Velocity Error Constant (K_V)	$K_V = \lim_{s\to 0} sG(s)H(s)$	$e_{ss} = \frac{A}{K_V}$
Acceleration Error Constant (K_A)	$K_A = \lim_{s\to 0} s^2 G(s)H(s)$	$e_{ss} = \frac{A}{K_A}$

STATIC ERROR COEFFICIENT METHOD

The general form of G(s)H(s) is

$$G(s)H(s) = \frac{K(1+T_1 s)(1+T_2 s)...(1+T_n s)}{s^j (1+T_a s)(1+T_b s)...(1+T_m s)}$$

Here, j = no. of poles at origin (s = 0)

Type 0

$$G(s)H(s) = \frac{K(1+T_1 s)(1+T_2 s)...(1+T_n s)}{(1+T_a s)(1+T_b s)...(1+T_m s)}$$

Here,

$$K_P = \lim_{s\to 0} G(s)H(s) = K$$

Therefore,

$$e_{ss} = \frac{A}{1+K}$$

Type 1

$$G(s)H(s) = \frac{K(1+T_1 s)(1+T_2 s)...(1+T_n s)}{s(1+T_a s)(1+T_b s)...(1+T_m s)}$$

Here,

$$K_V = \lim_{s\to 0} sG(s)H(s) = K$$

Therefore,

$$e_{ss} = \frac{A}{K}$$

Type 2

$$G(s)H(s) = \frac{K(1+T_1 s)(1+T_2 s)\ldots(1+T_n s)}{s^2(1+T_a s)(1+T_b s)\ldots(1+T_m s)}$$

Here,

$$K_A = \lim_{s \to 0} s^2 G(s)H(s) = K$$

Therefore,

$$e_{ss} = \frac{A}{K}$$

Steady-state error and error constant for different types of input are summarized as follows.

Type	Step input		Ramp input		Parabolic input	
	K_P	e_{ss}	K_V	e_{ss}	K_A	e_{ss}
Type 0	K	$\frac{A}{1+K}$	0	∞	0	∞
Type 1	∞	0	K	$\frac{A}{K}$	0	∞
Type 2	∞	0	∞	0	K	$\frac{A}{K}$

EFFECT OF POLE AND ZERO TO TRANSFER FUNCTION

(i) Addition of a pole to the Forward Path Transfer Function:-

(a) Increases the order of the system
(b) Increases the overshoot
(c) Reduces stability
(d) Increase rise time
(e) Reduces bandwidth

(ii) Addition of a pole to the Closed-Loop Transfer function:-

(a) Increases rise time
(b) Decreases overshoot

(iii) Addition of a zero to the Closed-Loop Transfer function:-

(a) Decreases rise time
(b) Increases overshoot

(iv) Addition of a zero to the Forward path Transfer function:-

(a) Added zero far away from imaginary axis – Overshoot large & damping is very poor
(b) When zero moves to the right – Overshoot reduce & damping improves
(c) When zero moves closer to the origin – Overshoot increases & damping improves

PROPORTIONAL CONTROLLER

The proportional controller produces an output, which is proportional to error signal.

$$u(t) \propto e(t)$$

$$\Rightarrow u(t) = K_P e(t)$$

Apply Laplace transform on both the sides -

$$U(s) = K_P E(s)$$

$$\frac{U(s)}{E(s)} = K_P$$

Therefore, the transfer function of the proportional controller is K_P.

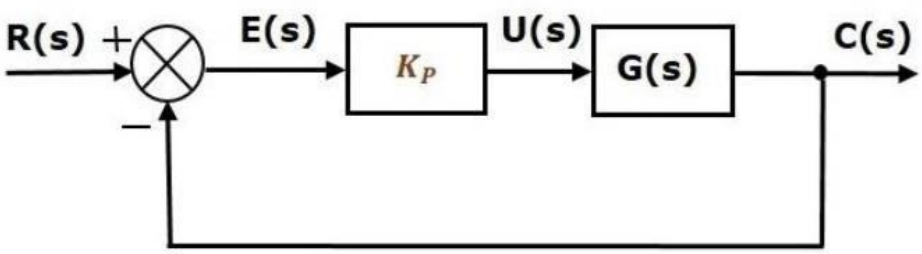
Where,

U(s) is the Laplace transform of the actuating signal u(t)

E(s) is the Laplace transform of the error signal e(t)

K_P is the proportionality constant

The block diagram of the unity negative feedback closed loop control system along with the proportional controller is shown in the following figure.



DERIVATIVE CONTROLLER

The derivative controller produces an output, which is derivative of the error signal.

$$u(t) = K_D \frac{de(t)}{dt}$$

Apply Laplace transform on both sides.

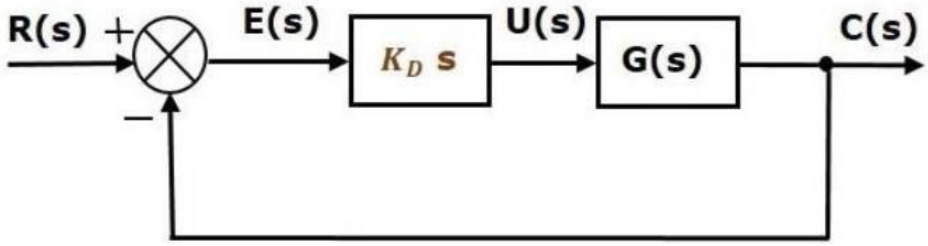
$$U(s) = K_D s E(s)$$

$$\frac{U(s)}{E(s)} = K_D s$$

Therefore, the transfer function of the derivative controller is $K_D s$

Where, K_D is the derivative constant.

The block diagram of the unity negative feedback closed loop control system along with the derivative controller is shown in the following figure.



INTEGRAL CONTROLLER

The integral controller produces an output, which is integral of the error signal.

$$u(t) = K_I \int e(t)dt$$

Apply Laplace transform on both the sides -

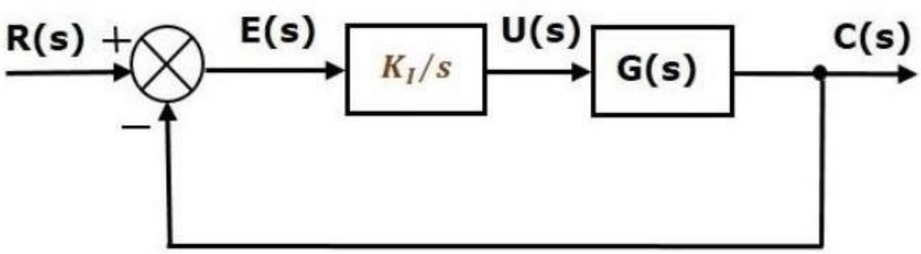
$$U(s) = \frac{K_I E(s)}{s}$$

$$\frac{U(s)}{E(s)} = \frac{K_I}{s}$$

Therefore, the transfer function of the integral controller is $\frac{K_I}{s}$.

Where, K_I is the integral constant.

The block diagram of the unity negative feedback closed loop control system along with the integral controller is shown in the following figure.



The integral controller is used to decrease the steady state error.

PROPORTIONAL DERIVATIVE (PD) CONTROLLER

The proportional derivative controller produces an output, which is the combination of the outputs of proportional and derivative controllers.

$$u(t) = K_P e(t) + K_D \frac{\mathrm{d}e(t)}{\mathrm{d}t}$$

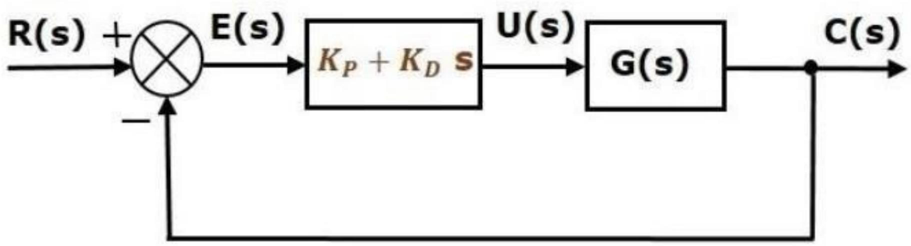
Apply Laplace transform on both sides -

$$U(s) = (K_P + K_D s)E(s)$$

$$\frac{U(s)}{E(s)} = K_P + K_D s$$

Therefore, the transfer function of the proportional derivative controller is $K_P + K_D s$.

The block diagram of the unity negative feedback closed loop control system along with the proportional derivative controller is shown in the following figure.



The proportional derivative controller is used to improve the stability of control system without affecting the steady state error.

PROPORTIONAL INTEGRAL (PI) CONTROLLER

The proportional integral controller produces an output, which is the combination of outputs of the proportional and integral controllers.

$$u(t) = K_P e(t) + K_I \int e(t)dt$$

Apply Laplace transform on both sides -

$$U(s) = \left(K_P + \frac{K_I}{s}\right)E(s)$$

$$\frac{U(s)}{E(s)} = K_P + \frac{K_I}{s}$$

Therefore, the transfer function of proportional integral controller is $K_P + \frac{K_I}{s}$.

The proportional integral controller is used to decrease the steady state error without affecting the stability of the control system.

PROPORTIONAL INTEGRAL DERIVATIVE (PID) CONTROLLER

The proportional integral derivative controller produces an output, which is the combination of the outputs of proportional, integral and derivative controllers.

$$u(t) = K_P e(t) + K_I \int e(t)dt + K_D \frac{de(t)}{dt}$$

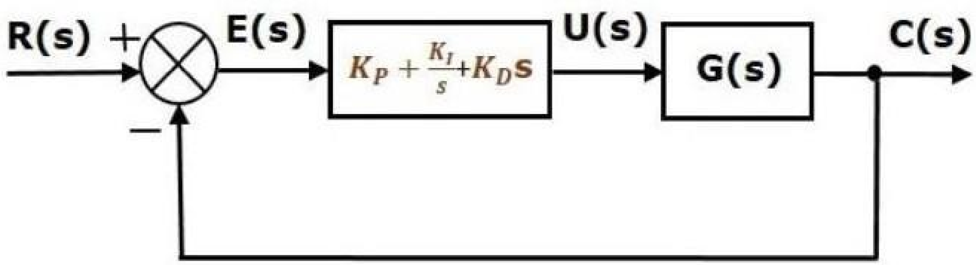
Apply Laplace transform on both sides -

$$U(s) = \left(K_P + \frac{K_I}{s} + K_D s\right) E(s)$$

$$\frac{U(s)}{E(s)} = K_P + \frac{K_I}{s} + K_D s$$

Therefore, the transfer function of the proportional integral derivative controller is $K_P + \frac{K_I}{s} + K_D s$.

The block diagram of the unity negative feedback closed loop control system along with the proportional integral derivative controller is shown in the following figure.



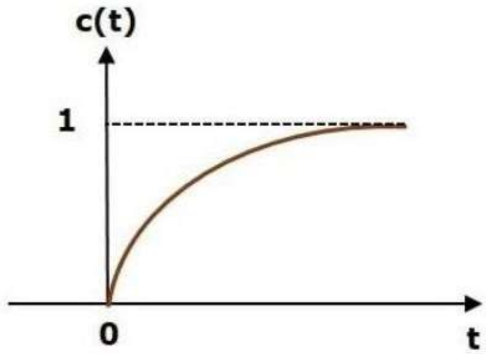
CHAPTER-06

ANALYSIS OF STABILITY BY ROOT LOCUS TECHNIQUE

STABILITY

A system is said to be stable, if its output is under control. Otherwise, it is said to be unstable. A **stable system** produces a bounded output for a given bounded input.

The following figure shows the response of a stable system.



This is the response of first order control system for unit step input. This response has the values between 0 and 1. So, it is bounded output. We know that the unit step signal has the value of one for all positive values of **t** including zero. So, it is bounded input. Therefore, the first order control system is stable since both the input and the output are bounded.

TYPES OF SYSTEMS BASED ON STABILITY

We can classify the systems based on stability as follows.

- Absolutely stable system
- Conditionally stable system
- Marginally stable system

Absolutely Stable System

If the system is stable for all the range of system component values, then it is known as the **absolutely stable system**. The open loop control system is absolutely stable if all the poles of the open loop transfer function present in left half of **'s' plane**. Similarly, the closed loop control system is absolutely stable if all the poles of the closed loop transfer function present in the left half of the 's' plane.

Conditionally Stable System

If the system is stable for a certain range of system component values, then it is known as **conditionally stable system**.

Marginally Stable System

If the system is stable by producing an output signal with constant amplitude and constant frequency of oscillations for bounded input, then it is known as **marginally stable system**. The open loop control system is marginally stable if any two poles of the open loop transfer function is present on the imaginary axis. Similarly, the closed loop control system is marginally stable if any two poles of the closed loop transfer function is present on the imaginary axis.

ROUTH-HURWITZ STABILITY CRITERION

Routh-Hurwitz stability criterion is having one necessary condition and one sufficient condition for stability. If any control system doesn't satisfy the necessary condition, then we can say that the control system is unstable. But, if the control system satisfies the necessary condition, then it may or may not be stable. So, the sufficient condition is helpful for knowing whether the control system is stable or not.

NECESSARY CONDITION FOR ROUTH-HURWITZ STABILITY

The necessary condition is that the coefficients of the characteristic polynomial should be positive. This implies that all the roots of the characteristic equation should have negative real parts.

Consider the characteristic equation of the order 'n' is -

$$a_0 s^n + a_1 s^{n-1} + a_2 s^{n-2} + \ldots + a_{n-1} s^1 + a_n s^0 = 0$$

Note that, there should not be any term missing in the **n**$^{\text{th}}$ order characteristic equation. This means that the **n**$^{\text{th}}$ order characteristic equation should not have any coefficient that is of zero value.

SUFFICIENT CONDITION FOR ROUTH-HURWITZ STABILITY

The sufficient condition is that all the elements of the first column of the Routh array should have the same sign. This means that all the elements of the first column of the Routh array should be either positive or negative.

ROUTH ARRAY METHOD

If all the roots of the characteristic equation exist to the left half of the 's' plane, then the control system is stable. If at least one root of the characteristic equation exists to the right half of the 's' plane, then the control system is unstable. So, we have to find the roots of the characteristic equation to know whether the control system is stable or unstable. But, it is difficult to find the roots of the characteristic equation as order increases.

So, to overcome this problem there we have the **Routh array method**. In this method, there is no need to calculate the roots of the characteristic equation. First formulate the Routh table and find the number of the sign changes in the first column of the Routh table. The number of sign changes in the first column of the Routh table gives the number of roots of characteristic equation that exist in the right half of the 's' plane and the control system is unstable.

Follow this procedure for forming the Routh table.

- Fill the first two rows of the Routh array with the coefficients of the characteristic polynomial as mentioned in the table below. Start with the coefficient of sn and continue up to the coefficient of s_0.
- Fill the remaining rows of the Routh array with the elements as mentioned in the table below. Continue this process till you get the first column element of **row** s_0.

Note − If any row elements of the Routh table have some common factor, then you can divide the row elements with that factor for the simplification will be easy.

The following table shows the Routh array of the n$^{\text{th}}$ order characteristic polynomial.

$$a_0s^n + a_1s^{n-1} + a_2s^{n-2} + \ldots + a_{n-1}s^1 + a_ns^0$$

s^n	a_0	a_2	a_4	a_6	$\ldots$	$\ldots$
s^{n-1}	a_1	a_3	a_5	a_7	$\ldots$	$\ldots$
s^{n-2}	$b_1 = \frac{a_1a_2 - a_3a_0}{a_1}$	$b_2 = \frac{a_1a_4 - a_5a_0}{a_1}$	$b_3 = \frac{a_1a_6 - a_7a_0}{a_1}$	$\ldots$	$\ldots$	$\ldots$
s^{n-3}	$c_1 = \frac{b_1a_3 - b_2a_1}{b_1}$	$c_2 = \frac{b_1a_55 - b_3a_1}{b_1}$	$\vdots$			
$\vdots$	$\vdots$	$\vdots$	$\vdots$			
s^1	$\vdots$	$\vdots$				
s^0	a_n					

Example

Let us find the stability of the control system having characteristic equation,

$$s^4 + 3s^3 + 3s^2 + 2s + 1 = 0$$

Step 1 − Verify the necessary condition for the Routh-Hurwitz stability.

All the coefficients of the characteristic polynomial,

$$s^4 + 3s^3 + 3s^2 + 2s + 1$$

are positive. So, the control system satisfies the necessary condition.

Step 2 − Form the Routh array for the given characteristic polynomial.

s^4	1	3	1
s^3	3	2	
s^2	$\frac{(3\times3)-(2\times1)}{3} = \frac{7}{3}$	$\frac{(3\times1)-(0\times1)}{3} = \frac{3}{3} = 1$	
s^1	$\frac{\left(\frac{7}{3}\times2\right)-(1\times3)}{\frac{7}{3}} = \frac{5}{7}$		
s^0	1		

Step 3 − Verify the sufficient condition for the Routh-Hurwitz stability.

All the elements of the first column of the Routh array are positive. There is no sign change in the first column of the Routh array. So, the control system is stable.

SPECIAL CASES OF ROUTH ARRAY

We may come across two types of situations, while forming the Routh table. It is difficult to complete the Routh table from these two situations.

The two special cases are −

- The first element of any row of the Routh's array is zero.
- All the elements of any row of the Routh's array are zero.

Let us now discuss how to overcome the difficulty in these two cases, one by one.

- **First Element of any row of the Routh's array is zero**

If any row of the Routh's array contains only the first element as zero and at least one of the remaining elements have non-zero value, then replace the first element with a small positive integer, ϵ. And then continue the process of completing the Routh's table. Now, find the number of sign changes in the first column of the Routh's table by substituting ϵ tends to zero.

- **All the Elements of any row of the Routh's array are zero**
  In this case, follow these two steps −

- Write the auxiliary equation, A(s) of the row, which is just above the row of zeros.
- Differentiate the auxiliary equation, A(s) with respect to s. fill the row of zeros with these coefficients.

ROOT LOCUS

The Root locus is the locus of the roots of the characteristic equation by varying system gain K from zero to infinity.

We know that, the characteristic equation of the closed loop control system is

$$1 + G(s)H(s) = 0$$

We can represent $G(s)H(s)$ as

$$G(s)H(s) = K\frac{N(s)}{D(s)}$$

Where,

- K represents the multiplying factor

- N(s) represents the numerator term having (factored) n^{th} order polynomial of 's'.
- D(s) represents the denominator term having (factored) m^{th} order polynomial of 's'.

Substitute, $G(s)H(s)$ value in the characteristic equation.

$$1 + k\frac{N(s)}{D(s)} = 0$$

$$\Rightarrow D(s) + KN(s) = 0$$

Case 1 − K = 0

If $K = 0$, then $D(s) = 0$.

That means, the closed loop poles are equal to open loop poles when K is zero.

Case 2 − K = ∞

Re-write the above characteristic equation as

$$K\left(\frac{1}{K} + \frac{N(s)}{D(s)}\right) = 0 \Rightarrow \frac{1}{K} + \frac{N(s)}{D(s)} = 0$$

Substitute, $K = \infty$ in the above equation.

$$\frac{1}{\infty} + \frac{N(s)}{D(s)} = 0 \Rightarrow \frac{N(s)}{D(s)} = 0 \Rightarrow N(s) = 0$$

If $K = \infty$, then $N(s) = 0$. It means the closed loop poles are equal to the open loop zeros when K is infinity.

From above two cases, we can conclude that the root locus branches start at open loop poles and end at open loop zeros.

ANGLE CONDITION AND MAGNITUDE CONDITION

The points on the root locus branches satisfy the angle condition. So, the angle condition is used to know whether the point exist on root locus branch or not. We can find the value of K for the points on the root locus branches by using magnitude condition. So, we can use the magnitude condition for the points, and this satisfies the angle condition.

Characteristic equation of closed loop control system is

$$1 + G(s)H(s) = 0$$

$$\Rightarrow G(s)H(s) = -1 + j0$$

The **phase angle** of $G(s)H(s)$ is

$$\angle G(s)H(s) = \tan^{-1}\left(\frac{0}{-1}\right) = (2n + 1)\pi$$

The **angle condition** is the point at which the angle of the open loop transfer function is an odd multiple of 180^0

Magnitude of G(s)H(s) is –

$$|G(s)H(s)| = \sqrt{(-1)^2 + 0^2} = 1$$

The magnitude condition is that the point (which satisfied the angle condition) at which the magnitude of the open loop transfer function is one.

THE ROOT LOCUS IS A GRAPHICAL REPRESENTATION IN S-DOMAIN AND IT IS SYMMETRICAL ABOUT THE REAL AXIS. Because the open loop poles and zeros exist in the s-domain having the values either as real or as complex conjugate pairs.

RULES FOR CONSTRUCTION OF ROOT LOCUS

Follow these rules for constructing a root locus.

Rule 1 − Locate the open loop poles and zeros in the‘s’ plane.

Rule 2 − Find the number of root locus branches.

We know that the root locus branches start at the open loop poles and end at open loop zeros. So, the number of root locus branches **N** is equal to the number of finite open loop poles **P** or the number of finite open loop zeros **Z**, whichever is greater.

Mathematically, we can write the number of root locus branches **N** as

$N=P$ if $P \geq Z$
$N=Z$ if $P<Z$

Rule 3 − Identify and draw the **real axis root locus branches**.

If the angle of the open loop transfer function at a point is an odd multiple of 180^0, then that point is on the root locus. If odd number of the open loop poles and zeros exist to the left side of a point on the real axis, then that point is on the root locus branch. Therefore, the branch of points which satisfies this condition is the real axis of the root locus branch.

Rule 4 − Find the centroid and the angle of asymptotes.

- If P=Z, then all the root locus branches start at finite open loop poles and end at finite open loop zeros.
- If P>Z, then Z number of root locus branches start at finite open loop poles and end at finite open loop zeros and P−Z number of root locus branches start at finite open loop poles and end at infinite open loop zeros.
- If P<Z , then P number of root locus branches start at finite open loop poles and end at finite open loop zeros and Z−P number of root locus branches start at infinite open loop poles and end at finite open loop zeros.

So, some of the root locus branches approach infinity, when P≠Z. Asymptotes give the direction of these root locus branches. The intersection point of asymptotes on the real axis is known as **centroid**.

We can calculate the **centroid α** by using this formula,

$$\alpha = \frac{\sum Real\ part\ of\ finite\ open\ loop\ poles - \sum Real\ part\ of\ finite\ open\ loop\ zeros}{P-Z}$$

The formula for the angle of **asymptotes θ** is

$$\theta = \frac{(2q+1)180^0}{P-Z}$$

Where,

$$q = 0, 1, 2, \dots, (P-Z)-1$$

Rule 5 − Find the intersection points of root locus branches with an imaginary axis.

We can calculate the point at which the root locus branch intersects the imaginary axis and the value of **K** at that point by using the Routh array method and special **case (ii)**.

- If all elements of any row of the Routh array are zero, then the root locus branch intersects the imaginary axis and vice-versa.

- Identify the row in such a way that if we make the first element as zero, then the elements of the entire row are zero. Find the value of **K** for this combination.

- Substitute this **K** value in the auxiliary equation. You will get the intersection point of the root locus branch with an imaginary axis.

Rule 6 − Find Break-away and Break-in points.

- If there exists a real axis root locus branch between two open loop poles, then there will be a **break-away point** in between these two open loop poles.
- If there exists a real axis root locus branch between two open loop zeros, then there will be a **break-in point** in between these two open loop zeros.

Note − Break-away and break-in points exist only on the real axis root locus branches.

Follow these steps to find break-away and break-in points.

- Write K in terms of s from the characteristic equation 1+G(s)H(s)=0.
- Differentiate K with respect to s and make it equal to zero. Substitute these values of s in the above equation.
- The values of s for which the K value is positive are the **break points**.

Rule 7 − Find the angle of departure and the angle of arrival.

The Angle of departure and the angle of arrival can be calculated at complex conjugate open loop poles and complex conjugate open loop zeros respectively.

The formula for the **angle of departure** ϕ_d is

$$\phi_d = 180^0 - \phi$$

The formula for the **angle of arrival** ϕ_a is

$$\phi_a = 180^0 + \phi$$

Where,

$$\phi = \sum \phi_P - \sum \phi_Z$$

Example

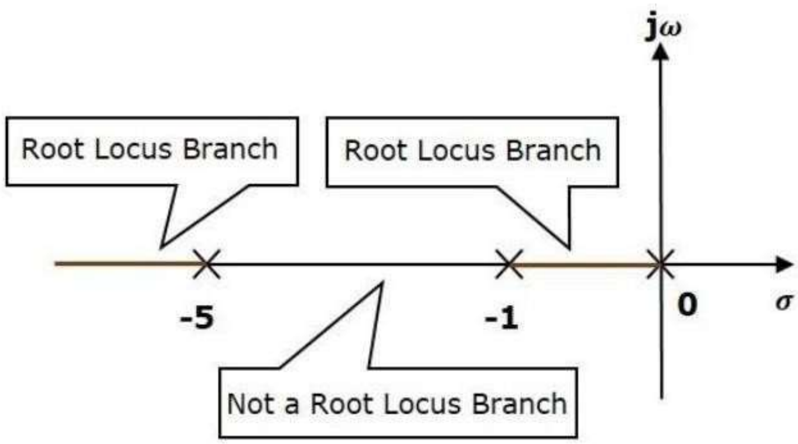
Let us now draw the root locus of the control system having open loop transfer function, $G(s)H(s) = \frac{K}{s(s+1)(s+5)}$

Step 1 − The given open loop transfer function has three poles at s = 0,

s = -1, s = -5. It doesn't have any zero. Therefore, the number of root locus branches is equal

to the number of poles of the open loop transfer function.

N=P=3

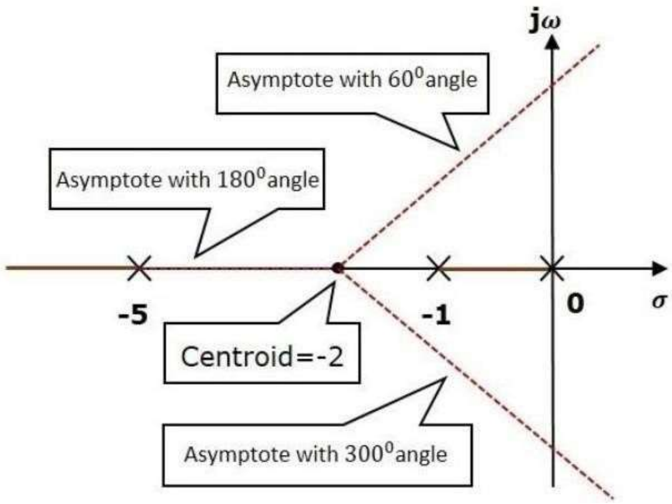


The three poles are located are shown in the above figure. The line segment between s=−1, and s=0 is one branch of root locus on real axis. And the other branch of the root locus on the real axis is the line segment to the left of s=−5.

Step 2 − We will get the values of the centroid and the angle of asymptotes by using the given formulae.

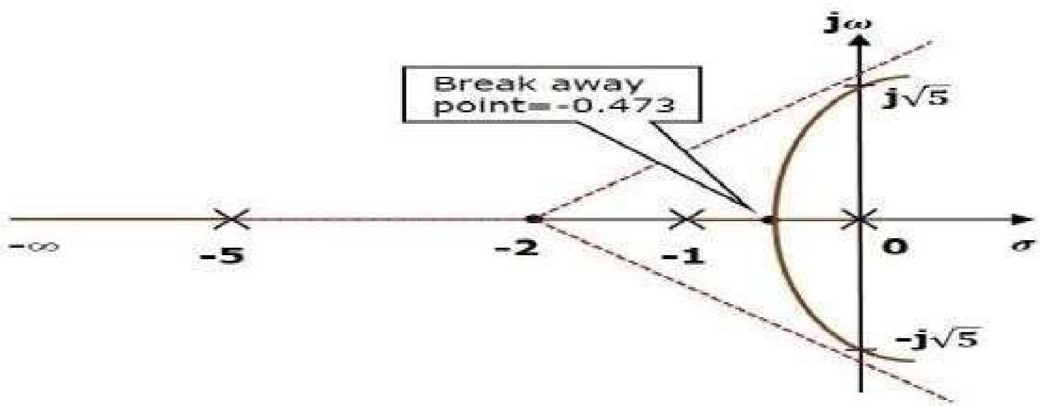
Centroid = $\frac{0-5-1}{3} = -2$

The angle of asymptotes are $\theta = 60^0, 180^0$ and 300^0.



Step 3 − Since two asymptotes have the angles of 60 and 300, two root locus branches intersect the imaginary axis. By using the Routh array method and special case(ii),
the root locus branches intersects the imaginary axis at $j\sqrt{5}$ and $-j\sqrt{5}$.

There will be one break-away point on the real axis root locus branch between the poles s =−1 and s=0. By following the procedure given for the calculation of break-away point, we will get it as s =−0.473.



EFFECTS OF ADDING OPEN LOOP POLES AND ZEROS ON ROOT LOCUS

The root locus can be shifted in's' **plane** by adding the open loop poles and the open loop zeros.

- If we include a pole in the open loop transfer function, then some of root locus branches will move towards right half of 's' plane. Because of this, the damping ratio δ decreases. Which implies, damped frequency ωd increases and the time domain specifications like delay time td, rise time tr and peak time tp decrease. But, it effects the system stability.

- If we include a zero in the open loop transfer function, then some of root locus branches will move towards left half of 's' plane. So, it will increase the control system stability. In this case, the damping ratio δ increases. Which implies, damped frequency ωd decreases and the time domain specifications like delay time td, rise time tr and peak time tp increase.

CHAPTER-07

FREQUENCY RESPONSE ANALYSIS

FREQUENCY RESPONSE

The response of a system can be partitioned into both the transient response and the steady state response. We can find the transient response by using Fourier integrals. The steady state response of a system for an input sinusoidal signal is known as the **frequency response**. In this chapter, we will focus only on the steady state response.

If a sinusoidal signal is applied as an input to a Linear Time-Invariant (LTI) system, then it produces the steady state output, which is also a sinusoidal signal. The input and output sinusoidal signals have the same frequency, but different amplitudes and phase angles. Let the input signal be

$$r(t) = A\sin(\omega_0 t)$$

The open loop transfer function will be −

$$G(s) = G(j\omega)$$

We can represent $G(j\omega)$ in terms of magnitude and phase as shown below.

$$G(j\omega) = |G(j\omega)|\angle G(j\omega)$$

Substitute, $\omega = \omega_0$ in the above equation.

$$G(j\omega_0) = |G(j\omega_0)|\angle G(j\omega_0)$$

The output signal is

$$c(t) = A|G(j\omega_0)|\sin(\omega_0 t + \angle G(j\omega_0))$$

- The **amplitude** of the output sinusoidal signal is obtained by multiplying the amplitude of the input sinusoidal signal and the magnitude of $G(j\omega)$ at $\omega = \omega_0$.
- The **phase** of the output sinusoidal signal is obtained by adding the phase of the input sinusoidal signal and the phase of $G(j\omega)$ at $\omega = \omega_0$.

Where,

A is the amplitude of the input sinusoidal signal.

ω_0 is angular frequency of the input sinusoidal signal.

We can write, angular frequency ω0 as shown below.

$\omega_0 = 2\pi f_0$

Here, f_0 is the frequency of the input sinusoidal signal. Similarly, you can follow the same procedure for closed loop control system

FREQUENCY DOMAIN SPECIFICATIONS

The frequency domain specifications are

- **Resonant peak**
- **Resonant frequency**
- **Bandwidth**.

Consider the transfer function of the second order closed control system as

$$T(s) = \frac{C(s)}{R(s)} = \frac{\omega_n^2}{s^2 + 2\delta\omega_n s + \omega_n^2}$$

Substitute, $s = j\omega$ in the above equation.

$$T(j\omega) = \frac{\omega_n^2}{(j\omega)^2 + 2\delta\omega_n (j\omega) + \omega_n^2}$$

$$\Rightarrow T(j\omega) = \frac{\omega_n^2}{-\omega^2 + 2j\delta\omega\omega_n + \omega_n^2} = \frac{\omega_n^2}{\omega_n^2 \left(1 - \frac{\omega^2}{\omega_n^2} + \frac{2j\delta\omega}{\omega_n}\right)}$$

$$\Rightarrow T(j\omega) = \frac{1}{\left(1 - \frac{\omega^2}{\omega_n^2}\right) + j\left(\frac{2\delta\omega}{\omega_n}\right)}$$

Let, $\frac{\omega}{\omega_n} = u$ Substitute this value in the above equation.

$$T(j\omega) = \frac{1}{(1 - u^2) + j(2\delta u)}$$

Magnitude of $T(j\omega)$ is -

$$M = |T(j\omega)| = \frac{1}{\sqrt{(1 - u^2)^2 + (2\delta u)^2}}$$

Phase of $T(j\omega)$ is -

$$\angle T(j\omega) = -tan^{-1}\left(\frac{2\delta u}{1 - u^2}\right)$$

Resonant Frequency

It is the frequency at which the magnitude of the frequency response has peak value for the first time. It is denoted by ω_r. At $\omega = \omega_r$, the first derivate of the magnitude of $T(j\omega)$ is zero.

Differentiate M with respect to u.

$$\frac{dM}{du} = -\frac{1}{2}\left[(1-u^2)^2 + (2\delta u)^2\right]^{\frac{-3}{2}} \left[2(1-u^2)(-2u) + 2(2\delta u)(2\delta)\right]$$

$$\Rightarrow \frac{dM}{du} = -\frac{1}{2}\left[(1-u^2)^2 + (2\delta u)^2\right]^{\frac{-3}{2}} \left[4u(u^2 - 1 + 2\delta^2)\right]$$

Substitute, $u = u_r$ and $\frac{dM}{du} == 0$ in the above equation.

$$0 = -\frac{1}{2}\left[(1-u_r^2)^2 + (2\delta u_r)^2\right]^{-\frac{3}{2}} \left[4u_r(u_r^2 - 1 + 2\delta^2)\right]$$

$$\Rightarrow 4u_r(u_r^2 - 1 + 2\delta^2) = 0$$

$$\Rightarrow u_r^2 - 1 + 2\delta^2 = 0$$

$$\Rightarrow u_r^2 = 1 - 2\delta^2$$

$$\Rightarrow u_r = \sqrt{1-2\delta^2}$$

Substitute, $u_r = \frac{\omega_r}{\omega_n}$ in the above equation.

$$\frac{\omega_r}{\omega_n} = \sqrt{1-2\delta^2}$$

$$\Rightarrow \omega_r = \omega_n\sqrt{1-2\delta^2}$$

Resonant Peak:

It is the peak (maximum) value of the magnitude of T(jω). It is denoted by Mr. At $u=u_r$, the Magnitude of T(jω) is -

$$M_r = \frac{1}{\sqrt{(1-u_r^2)^2 + (2\delta u_r)^2}}$$

Substitute, $u_r = \sqrt{1-2\delta^2}$ and $1 - u_r^2 = 2\delta^2$ in the above equation.

$$M_r = \frac{1}{\sqrt{(2\delta^2)^2 + (2\delta\sqrt{1-2\delta^2})^2}}$$

$$\Rightarrow M_r = \frac{1}{2\delta\sqrt{1-\delta^2}}$$

Resonant peak in frequency response corresponds to the peak overshoot in the time domain transient response for certain values of damping ratio δ. So, the resonant peak and peak overshoot are correlated to each other.

Bandwidth:

It is the range of frequencies over which, the magnitude of T(jω) drops to 70.7% from its zero frequency value.

At ω=0, the value of u will be zero.

Substitute, u=0 in M.

$$M = \frac{1}{\sqrt{(1-0^2)^2+(2\delta(0))^2}} = 1$$

Therefore, the magnitude of T(jω) is one at ω=0

At 3-dB frequency, the magnitude of T(jω) will be 70.7% of magnitude of T(jω)) at ω=0

i.e., at $\omega = \omega_B, M = 0.707(1) = \frac{1}{\sqrt{2}}$

$$\Rightarrow M = \frac{1}{\sqrt{2}} = \frac{1}{\sqrt{(1-u_b^2)^2+(2\delta u_b)^2}}$$

$$\Rightarrow 2 = (1-u_b^2)^2 + (2\delta)^2 u_b^2$$

Let, $u_b^2 = x$

$$\Rightarrow 2 = (1-x)^2 + (2\delta)^2 x$$

$$\Rightarrow x^2 + (4\delta^2 - 2)x - 1 = 0$$

$$\Rightarrow x = \frac{-(4\delta^2-2) \pm \sqrt{(4\delta^2-2)^2+4}}{2}$$

Consider only the positive value of x.

$$x = 1 - 2\delta^2 + \sqrt{(2\delta^2-1)^2+1}$$

$$\Rightarrow x = 1 - 2\delta^2 + \sqrt{(2-4\delta^2+4\delta^4)}$$

Substitute, $x = u_b^2 = \frac{\omega_b^2}{\omega_n^2}$

$$\frac{\omega_b^2}{\omega_n^2} = 1 - 2\delta^2 + \sqrt{(2-4\delta^2+4\delta^4)}$$

$$\Rightarrow \omega_b = \omega_n \sqrt{1 - 2\delta^2 + \sqrt{(2-4\delta^2+4\delta^4)}}$$

Bandwidth ω_b in the frequency response is inversely proportional to the rise time tr in the time domain transient response.

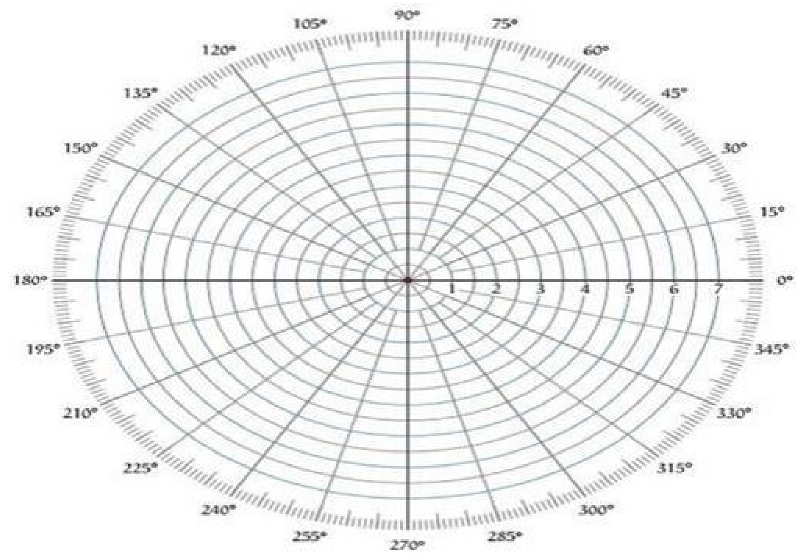
POLAR PLOTS

Polar plot is a plot which can be drawn between magnitude and phase. Here, the magnitudes are represented by normal values only.

The polar form of $G(j\omega)H(j\omega)$ is

$$G(j\omega)H(j\omega) = |G(j\omega)H(j\omega)| \angle G(j\omega)H(j\omega)$$

The **Polar plot** is a plot, which can be drawn between the magnitude and the phase angle of $G(j\omega)H(j\omega)$ by varying ω from zero to ∞. The polar graph sheet is shown in the following figure.



This graph sheet consists of concentric circles and radial lines. The **concentric circles** and the **radial lines** represent the magnitudes and phase angles respectively. These angles are represented by positive values in anti-clock wise direction. Similarly, we can represent angles with negative values in clockwise direction. For example, the angle 270^0 in anti-clock wise direction is equal to the angle -90^0 in clockwise direction.

RULES FOR DRAWING POLAR PLOTS

Follow these rules for plotting the polar plots.

- Substitute, s=jω in the open loop transfer function.
- Write the expressions for magnitude and the phase of G(jω)H(jω)
- Find the starting magnitude and the phase of G(jω)H(jω) by substituting ω=0. So, the polar plot starts with this magnitude and the phase angle.
- Find the ending magnitude and the phase of G(jω)H(jω) by substituting ω=∞ So, the polar plot ends with this magnitude and the phase angle.
- Check whether the polar plot intersects the real axis, by making the imaginary term of G(jω)H(jω) equal to zero and find the value(s) of ω.
- Check whether the polar plot intersects the imaginary axis, by making real term of G(jω)H(jω) equal to zero and find the value(s) of ω.
- For drawing polar plot more clearly, find the magnitude and phase of G(jω)H(jω) by considering the other value(s) of ω.

Example:

Consider the open loop transfer function of a closed loop control system.

$$G(s)H(s) = \frac{5}{s(s+1)(s+2)}$$

Let us draw the polar plot for this control system using the above rules.

Step 1 − Substitute, $s = j\omega$ in the open loop transfer function.

$$G(j\omega)H(j\omega) = \frac{5}{j\omega(j\omega+1)(j\omega+2)}$$

The magnitude of the open loop transfer function is

$$M = \frac{5}{\omega(\sqrt{\omega^2+1})(\sqrt{\omega^2+4})}$$

The phase angle of the open loop transfer function is

$$\phi = -90^0 - \tan^{-1}\omega - \tan^{-1}\frac{\omega}{2}$$

Frequency (rad/sec)	Magnitude	Phase angle(degrees)
0	∞	-90 or 270
∞	0	-270 or 90

So, the polar plot starts at $(\infty,-90^0)$ and ends at $(0,-270^0)$. The first and the second terms within the brackets indicate the magnitude and phase angle respectively.

Step 3 − Based on the starting and the ending polar co-ordinates, this polar plot will intersect the negative real axis. The phase angle corresponding to the negative real axis is -180^0 or 180^0. So, by equating the phase angle of the open loop transfer function to either -180^0 or 180^0, we will get the ω value as $\sqrt{2}$.

By substituting $\omega=\sqrt{2}$ in the magnitude of the open loop transfer function, we will get M=0.83. Therefore, the polar plot intersects the negative real axis when $\omega=\sqrt{2}$ and the polar coordinate is $(0.83,-180^0)$.

So, we can draw the polar plot with the above information.

BODE PLOTS

The Bode plot or the Bode diagram consists of two plots −

- Magnitude plot
- Phase plot

In both the plots, x-axis represents angular frequency (logarithmic scale). Whereas, yaxis represents the magnitude (linear scale) of open loop transfer function in the magnitude plot and the phase angle (linear scale) of the open loop transfer function in the phase plot.

The **magnitude** of the open loop transfer function in dB is -

$$M = 20\ \log|G(j\omega)H(j\omega)|$$

The **phase angle** of the open loop transfer function in degrees is -

$$\phi = \angle G(j\omega)H(j\omega)$$

Basic of Bode Plots:

The following table shows the slope, magnitude and the phase angle values of the terms present in the open loop transfer function. This data is useful while drawing the Bode plots.

Type of term	G(jω)H(jω)	Slope(dB/dec)	Magnitude (dB)	Phase angle(degrees)
Constant	K	0	$20 \log K$	0
Zero at origin	$j\omega$	20	$20 \log \omega$	90
'n' zeros at origin	$(j\omega)^n$	$20\,n$	$20\,n \log \omega$	$90\,n$
Pole at origin	$\frac{1}{j\omega}$	-20	$-20 \log \omega$	$-90\ or\ 270$
'n' poles at origin	$\frac{1}{(j\omega)^n}$	$-20\,n$	$-20\,n \log \omega$	$-90\,n\ or\ 270\,n$
Simple zero	$1 + j\omega r$	20	$0\ for\ \omega < \frac{1}{r}$ $20\ \log \omega r\ for\ \omega > \frac{1}{r}$	$0\ for\ \omega < \frac{1}{r}$ $90\ for\ \omega > \frac{1}{r}$
Simple pole	$\frac{1}{1+j\omega r}$	-20	$0\ for\ \omega < \frac{1}{r}$ $-20\ \log \omega r\ for\ \omega > \frac{1}{r}$	$0\ for\ \omega < \frac{1}{r}$ $-90\ or\ 270\ for\ \omega > \frac{1}{r}$
Second order derivative term	$\omega_n^2 \left(1 - \frac{\omega^2}{\omega_n^2} + \frac{2\delta j\omega}{\omega_n}\right)$	40	$40\ \log\ \omega_n\ for\ \omega < \omega_n$ $20\ \log\ (2\delta\omega_n^2)\ for\ \omega = \omega_n$ $40\ \log\ \omega\ for\ \omega > \omega_n$	$0\ for\ \omega < \omega_n$ $90\ for\ \omega = \omega_n$ $180\ for\ \omega > \omega_n$
Second order integral term	$\frac{1}{\omega_n^2\left(1-\frac{\omega^2}{\omega_n^2}+\frac{2\delta j\omega}{\omega_n}\right)}$	-40	$-40\ \log\ \omega_n\ for\ \omega < \omega_n$ $-20\ \log\ (2\delta\omega_n^2)\ for\ \omega = \omega_n$ $-40\ \log\ \omega\ for\ \omega > \omega_n$	$-0\ for\ \omega < \omega_n$ $-90\ for\ \omega = \omega_n$ $-180\ for\ \omega > \omega_n$

Case-1:

Consider the open loop transfer function $G(s)H(s) = K$.

Magnitude $M = 20\ \log K$ dB

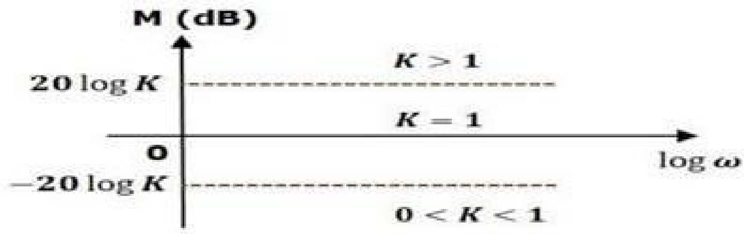
Phase angle $\phi = 0$ degrees

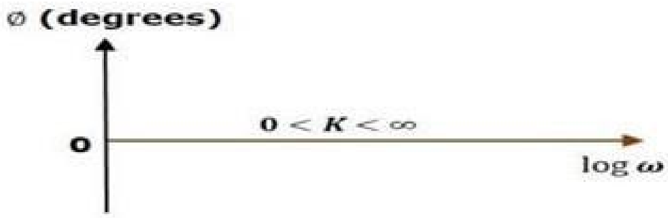
If $K = 1$, then magnitude is 0 dB.

If $K > 1$, then magnitude will be positive.

If $K < 1$, then magnitude will be negative.

The following figure shows the corresponding Bode plot.





The magnitude plot is a horizontal line, which is independent of frequency. The 0 dB line itself - is the magnitude plot when the value of K is one. For the positive values of K, the horizontal line will shift 20logK dB above the 0 dB line. For the negative values of K, the horizontal line will shift 20logK dB below the 0 dB line. The Zero degrees line itself is the phase plot for all the positive values of K.

Case-2:

Consider the open loop transfer function **G(s)H(s)= S**

Magnitude M=20logω dB

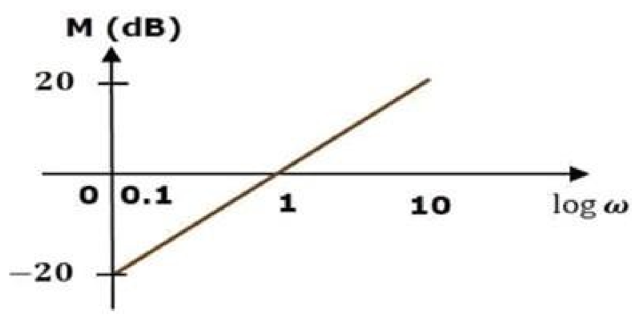
Phase angle $\phi=90^0$

At ω=0.1rad/sec, the magnitude is -20 dB.

At ω=1rad/sec, the magnitude is 0 dB.

At ω=10 rad/sec, the magnitude is 20 dB.

The following figure shows the corresponding Bode plot.



The magnitude plot is a line, which is having a slope of 20 dB/dec. This line started at ω=0.1rad/sec having a magnitude of -20 dB and it continues on the same slope. It is touching 0 dB line at ω=1 rad/sec. In this case, the phase plot is 90^0 line.

Case-3:

Consider the open loop transfer function **G(s) H(s)=1+sτ**.

Magnitude = $\sqrt{1+(s\tau)}^{2}$

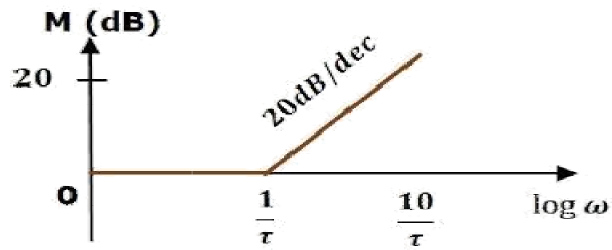
Phase angle =

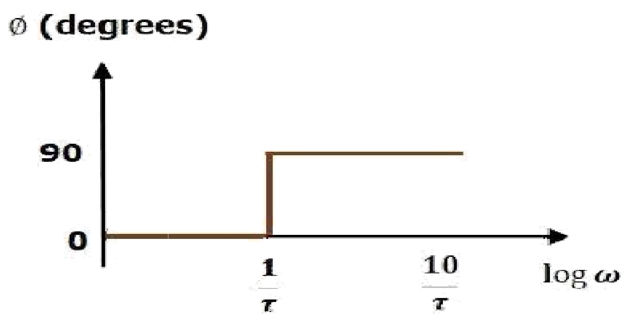
$\phi = \tan^{-1}\omega\tau$ degrees

For $\omega < \frac{1}{\tau}$, the magnitude is 0 dB and phase angle is 0 degrees.

For $\omega > \frac{1}{\tau}$, the magnitude is 20logωτ dB and phase angle is 90^0.

The following figure shows the corresponding Bode plot





The magnitude plot is having magnitude of 0 dB up to ω=1τ rad/sec. From ω=1τ rad/sec, it is having a slope of 20 dB/decade. In this case, the phase plot is having phase angle of 0 degrees up to ω=1τ rad/sec and from here, it is having phase angle of 90^0. This Bode plot is called the **asymptotic Bode plot**. As the magnitude and the phase plots are represented with straight lines, the Exact Bode plots resemble the asymptotic Bode plots. The only difference is that the Exact Bode plots will have simple curves instead of straight lines.

<u>RULES FOR CONSTRUCTION OF BODE PLOTS:</u>

Follow these rules while constructing a Bode plot.

- Represent the open loop transfer function in the standard time constant form.
- Substitute, s=jω in the above equation.
- Find the corner frequencies and arrange them in ascending order.
- Consider the starting frequency of the Bode plot as $1/10^{th}$ of the minimum corner frequency or 0.1 rad/sec whichever is smaller value and draw the Bode plot upto 10 times maximum corner frequency.
- Draw the magnitude plots for each term and combine these plots properly.
- Draw the phase plots for each term and combine these plots properly.

STABILITY ANALYSIS USING BODE PLOTS

From the Bode plots, we can say whether the control system is stable, marginally stable or unstable based on the values of these parameters.

- Gain cross over frequency and phase cross over frequency
- Gain margin and phase margin

Phase Cross over Frequency:
The frequency at which the phase plot is having the phase of -180^0 is known as **phase cross over frequency**. It is denoted by ω_{pc}. The unit of phase cross over frequency is **rad/sec**.

Gain Cross over Frequency:
The frequency at which the magnitude plot is having the magnitude of zero dB is known as **gain cross over frequency**. It is denoted by ω_{gc}. The unit of gain cross over frequency is **rad/sec**.

The stability of the control system based on the relation between the phase cross over frequency and the gain cross over frequency is listed below.

- If the phase cross over frequency ω_{pc} is greater than the gain cross over frequency ω_{gc}, then the control system is **stable**.
- If the phase cross over frequency ω_{pc} is equal to the gain cross over frequency ω_{gc}, then the control system is **marginally stable**.
- If the phase cross over frequency ω_{pc} is less than the gain cross over frequency ω_{gc}, then the control system is **unstable**.

Gain Margin:
Gain margin GM is equal to negative of the magnitude in dB at phase cross over frequency.

$$\mathbf{GM = -\,20\log(M_{pc})}$$

Where, Mpc is the magnitude at phase cross over frequency. The unit of gain margin (GM) is **dB**.

Phase Margin:
The formula for phase margin PM is $\mathbf{PM=180^0+\phi_{gc}}$
Where, ϕ_{gc} is the phase angle at gain cross over frequency. The unit of phase margin is **degrees**.

****The stability of the control system based on the relation between gain margin and phase margin is listed below.

- If both the gain margin GM and the phase margin PM are positive, then the control system is **stable**.
- If both the gain margin GM and the phase margin PM are equal to zero, then the control system is **marginally stable**.
- If the gain margin GM and / or the phase margin PM are/is negative, then the control system is **unstable**.

Example: Sketch the Bode plot for the Transfer function

$$G(s) = \frac{1000}{(1+0.1s)(1+0.001s)}$$

Determine the a) Phase Margin

b) Gain Margin

c) Stability of the System

Solution: Step-1 Put $s = j\omega$

$$G(j\omega) = \frac{1000}{(1+j0.1\omega)(1+j0.001\omega)}$$

The given transfer function is of type '0' system. Therefore the initial slope of the Bode plots 0 db/decade. The starting point is given by.

$$20 \log_{10} K = 20 \log_{10} 1000 = 60 \text{ db}$$

Corner frequencies $\omega_1 = \frac{1}{0.1} = 10$ rad/sec.

$\omega_2 = \frac{1}{0.001} = 1000$ rad/sec.

Step 2 : Mark the starting point 60 db on y-axis and draw a line of slop 0 db/decade up to first corner frequency.

Step 3 : From first corner frequency to second corner frequency draw a line with slope (0 – 20) = -20 db/decade).

Step 4 : From second corner frequency to next corner frequency (if given) draw a line having the slope -20 + (-20) = -40 db/decade.

Step 5 : The magnitude plot is complete and now draw the phase plot by calculating the phase at different frequencies (as given in table).

Step 6: From the bode plot

From the point of intersection of magnitude curve with 0 db axis draw a line on phase curve. This line cuts the phase curve at - 154°

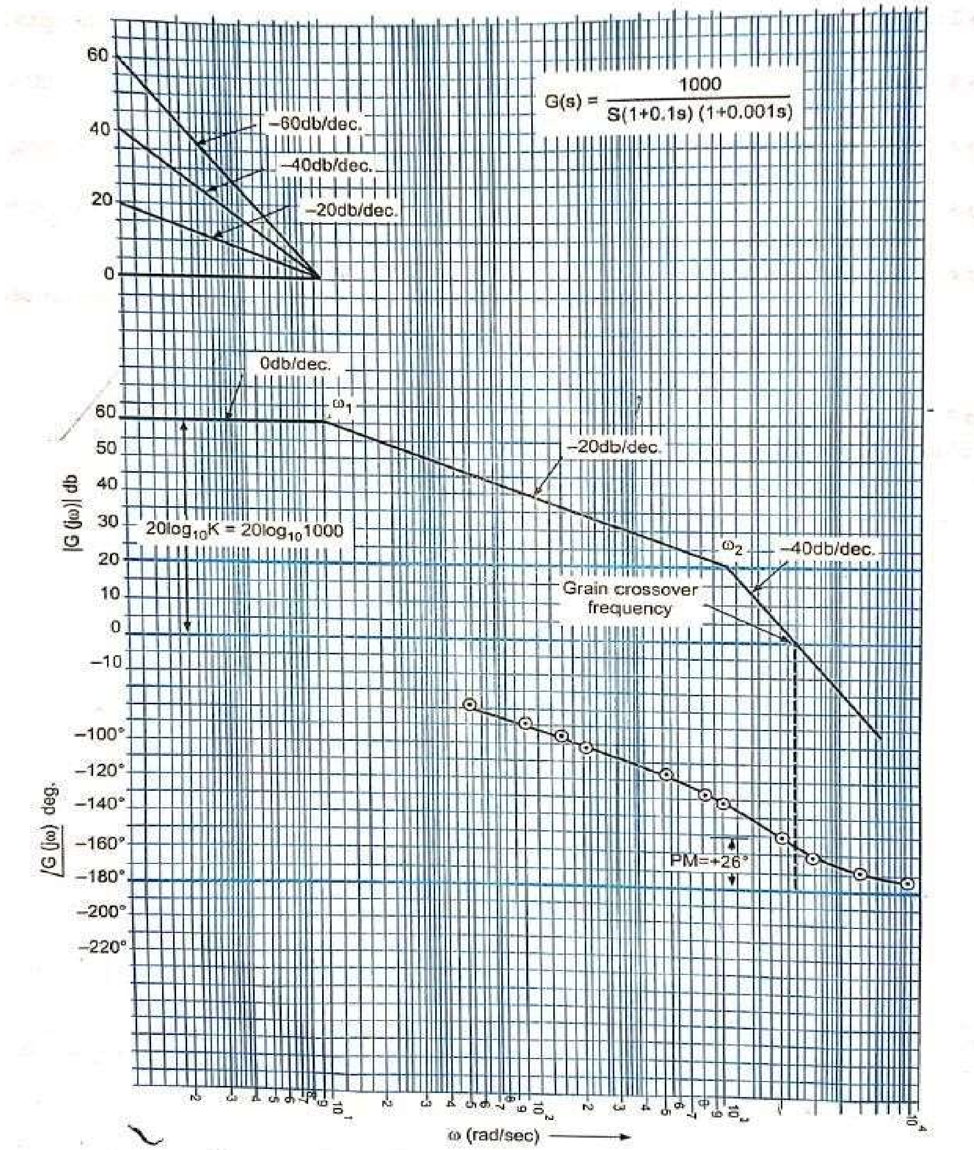
P.M = -154-(-180)

= +26°

Step 7: Gain margin G.M $=\infty$

Since, P.M = + 26° and gain margin $=\infty$, the system is inherently stable.

ω	- Arg $(1 + j0.1\omega)$ $-\tan^{-1}(0.1\,\omega)$	-Arg $(1 + j0.001\omega)$ $-\tan^{-1}(0.001\,\omega)$	Resultant
50	-78.6°	-2.86°	- 81.46°
100	- 84.2°	-5.7°	- 90°
150	- 86.2°	- 8.5°	- 94°
200	- 87.13°	- 11.3°	- 98°
500	- 88.85°	-26.56°	-115.4°
800	- 89.28°	- 38.65°	-127.93°
1000	-89.48°	-45°	-134.42°
2000	- 89.72°	- 63.43°	-153.15°
3000	- 89.8°	- 71.56°	-161.36°
5000	- 89.88°	- 78.69°	-168.57°
8000	- 89.92°	- 82.87°	-172.79°

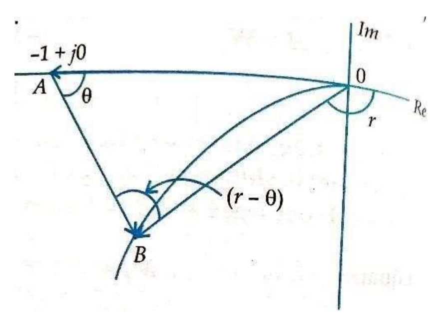


CLOSED LOOP FREQUENCY RESPONSE

Consider transfer function $\frac{C(s)}{R(s)} = \frac{G(s)}{1+G(s)H(s)}$

For unity feedback H(s) =1

$$\frac{C(s)}{R(s)} = \frac{G(s)}{1+G(s)}$$



Put s= jω

$$\frac{C(j\omega)}{R(j\omega)} = \frac{G(j\omega)}{1+G(j\omega)}$$

From figure

$$\vec{OB} = G(j\omega)$$

$$\vec{OA} = -1$$

$$\vec{AB} = \vec{OB} - \vec{OA} = G(j\omega) - (-1)$$

$$\vec{AB} = 1 + G(j\omega)$$

From above equation

$$\left|\frac{C(j\omega)}{R(j\omega)}\right| = M(\omega) = \frac{\overrightarrow{OB}}{\overrightarrow{AB}}$$

$$\angle\frac{C(j\omega)}{R(j\omega)} = \frac{\angle\overrightarrow{OB}}{\angle\overrightarrow{AB}} = \frac{r}{\theta} = r - \theta$$

$$\therefore \quad \frac{C(j\omega)}{R(j\omega)} = M(\omega)\, e^{j\,\phi(\omega)}$$

where $M(j\omega)$ is the magnitude and $\phi(\omega) = r - \theta$

Frequency response consists of 2 parts: (1) magnitude (2) phase angle. Both can be plotted against different values of ω.

CHAPTER-08

NYQUIST PLOT

NYQUIST PLOT

Nyquist plots are the continuation of polar plots for finding the stability of the closed loop control systems by varying ω from $-\infty$ to ∞. That means, Nyquist plots are used to draw the complete frequency response of the open loop transfer function.

NYQUIST STABILITY CRITERION

The Nyquist stability criterion works on the **principle of argument**. It states that if there are P poles and Z zeros are enclosed by the 's' plane closed path, then the corresponding G(s)H(s) plane must encircle the origin $P-Z$ times. So, we can write the number of encirclements N as,

$N=P-Z$

- ❖ If the enclosed's' plane closed path contains only poles, then the direction of the encirclement in the G(s)H(s) plane will be opposite to the direction of the enclosed closed path in the 's' plane.
- ❖ If the enclosed's' plane closed path contains only zeros, then the direction of the encirclement in the G(s)H(s) plane will be in the same direction as that of the enclosed closed path in the 's' plane.

Let us now apply the principle of argument to the entire right half of the's' plane by selecting it as a closed path. This selected path is called the **Nyquist** contour.

We know that the closed loop control system is stable if all the poles of the closed loop transfer function are in the left half of the's' plane. So, the poles of the closed loop transfer function are nothing but the roots of the characteristic equation. As the order of the characteristic equation increases, it is difficult to find the roots. So, let us correlate these roots of the characteristic equation as follows.

- The Poles of the characteristic equation are same as that of the poles of the open loop transfer function.
- The zeros of the characteristic equation are same as that of the poles of the closed loop transfer function.

We know that the open loop control system is stable if there is no open loop pole in the right half of the 's' plane. i.e. $P=0 \Rightarrow N=-Z$

We know that the closed loop control system is stable if there is no closed loop pole in the right half of the 's' plane. i.e. $Z=0 \Rightarrow N=P$

Nyquist stability criterion states the number of encirclements about the critical point (1+j0) must be equal to the poles of characteristic equation, which is nothing but the poles of the open loop transfer function in the right half of the 's' plane. The shift in origin to (1+j0) gives the characteristic equation plane.

RULES FOR DRAWING NYQUIST PLOTS

Follow these rules for plotting the Nyquist plots.

- Locate the poles and zeros of open loop transfer function G(s)H(s) in 's' plane.
- Draw the polar plot by varying ω from zero to infinity. If pole or zero present at s = 0, then varying ω from 0+ to infinity for drawing polar plot.
- Draw the mirror image of above polar plot for values of ω ranging from $-\infty$ to zero (0^- if any pole or zero present at s=0).
- The number of infinite radius half circles will be equal to the number of poles or zeros at origin. The infinite radius half circle will start at the point where the mirror image of the polar plot ends. And this infinite radius half circle will end at the point where the polar plot starts.

After drawing the Nyquist plot, we can find the stability of the closed loop control system using the Nyquist stability criterion. If the critical point (-1+j0) lies outside the encirclement, then the closed loop control system is absolutely stable.

STABILITY ANALYSIS USING NYQUIST PLOTS

From the Nyquist plots, we can identify whether the control system is stable, marginally stable or unstable based on the values of these parameters.

- Gain cross over frequency and phase cross over frequency
- Gain margin and phase margin

Phase Cross over Frequency

The frequency at which the Nyquist plot intersects the negative real axis (phase angle is 180^0) is known as the **phase cross over frequency**. It is denoted by ω_{pc}.

Gain Cross over Frequency

The frequency at which the Nyquist plot is having the magnitude of one is known as the **gain cross over frequency**. It is denoted by ω_{gc}.

The stability of the control system based on the relation between phase cross over frequency and gain cross over frequency is listed below.

- If the phase cross over frequency ω_{pc} is greater than the gain cross over frequency ω_{gc}, then the control system is **stable**.
- If the phase cross over frequency ω_{pc} is equal to the gain cross over frequency ω_{gc}, then the control system is **marginally stable**.
- If phase cross over frequency ω_{pc} is less than gain cross over frequency ω_{gc}, then the control system is **unstable**.

Gain Margin

The gain margin GM is equal to the reciprocal of the magnitude of the Nyquist plot at the phase cross over frequency.

$$GM = \frac{1}{M_{pc}}$$

Where, Mpc is the magnitude in normal scale at the phase cross over frequency.

Phase Margin

The phase margin PM is equal to the sum of 180^0 and the phase angle at the gain cross over frequency.

$$PM = 180^0 + \phi_{gc}$$

Where, ϕ_{gc} is the phase angle at the gain cross over frequency.

The stability of the control system based on the relation between the gain margin and the phase margin is listed below.

- ❖ If the gain margin GM is greater than one and the phase margin PM is positive, then the control system is **stable**.
- ❖ If the gain margin GMs equal to one and the phase margin PM is zero degrees, then the control system is **marginally stable**.
- ❖ If the gain margin GM is less than one and / or the phase margin PM is negative, then the control system is **unstable**.

Example:- Draw the nyquist plot and assess the stability of the closed loop system whose open loop transfer function is G(s) H(s) = $\dfrac{K}{s(2s+1)}$

$$G(j\omega)H(j\omega) = \frac{K}{j\omega(2j\omega+1)} = \frac{K(1-j2\omega)}{j\omega(1+j2\omega)(1-j2\omega)}$$

$$= \frac{K(1-j2\omega)}{j\omega(1+4\omega^2)} = -\frac{2K}{1+4\omega^2} - j\frac{K}{\omega(1+4\omega^2)}$$

Along the segment (C_1) of the Nyquist contour on the $j\omega$-axis, s varies from $-j\infty$ to $+j\infty$.

At $\omega = -\infty$,

$$G(j\omega)H(j\omega) = -0 + j0$$

At $\omega = 0^-$,

$$G(j\omega)H(j\omega) = -2K + j\infty$$

At $\omega = 0^+$,

$$G(j\omega)H(j\omega) = -2K - j\infty$$

At $\omega = +\infty$,

$$G(j\omega)H(j\omega) = -0 - j0$$

The infinitesimally small semicircular indent around the pole at the origin of Figure represented by $s = \underset{\in \to 0}{\in e^{j\theta}}$ (θ varying from $-\pi/2°$ through 0° to $+\pi/2°$) maps into

$$\underset{\in \to 0}{\text{Lt}} [K/(\in e^{j\theta})(2\in e^{j\theta} + 1)] = \underset{\in \to 0}{\text{Lt}} K/\in e^{j\theta} = \underset{\in \to 0}{\text{Lt}} (K/\in)e^{-j\theta} = \infty e^{-j\theta}$$

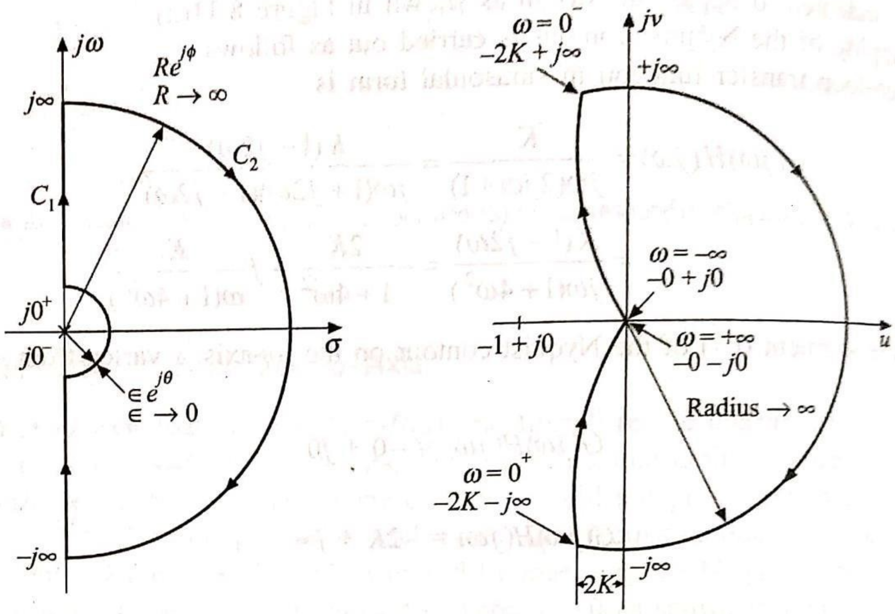
$-\theta$ varies from +90° → 0° → −90°. Thus, the infinitesimal semi-circular indent around the origin of the s-plane maps into a semicircular arc of infinite radius in $G(s)H(s)$ plane extending from +90° through 0° to −90°.

The infinite semicircular arc of the Nyquist contour (segment C_2) of Figure represented by $s = \underset{R \to \infty}{Re^{j\phi}}$ (ϕ varying from +90° through 0° to −90°) is mapped into

$$\underset{R \to \infty}{\text{Lt}} [K/(Re^{j\phi})(2Re^{j\phi} + 1)] = \underset{R \to \infty}{\text{Lt}} K/(2R^2e^{j2\phi}) = 0e^{-j2\phi}$$

that is, the origin of the $G(s)H(s)$ plane. The $G(s)H(s)$ locus thus turns at the origin with zero radius from −180° through 0° to +180°.

The given open-loop system has no poles in the right-half of the s-plane, i.e. $P = 0$. So as per the Nyquist stability criterion, for the closed-loop system to be stable, the Nyquist plot of $G(j\omega)H(j\omega)$ must not encircle the $(-1 + j0)$ point. Since the actual Nyquist plot in Figure 8.11(b) does not encircle the $(-1 + j0)$ point, the closed-loop system is always stable.



EFFECT OF ADDITION OF POLES & ZEROS TO G(s)H(s) ON THE SHAPE OF NYQUIST PLOT

a) **Addition of poles at s=0:-** It will affect the stability of the closed loop system adversely. A system that has a loop transfer function with more than one pole at s=0 is likely to be unstable or difficult to stabilize.

b) **Addition of finite non zero pole:-** It shifts the phase of nyquist plot by -90° at ω= ∞. The stability is adversely affected.

c) **Addition of a Zero: -** The effect of addition of zero is to rotate the nyquist plot by 90° in the counter clockwise direction without effecting the value at ω = 0. So it has the effect of reducing the overshoot & the general effect of stabilization.

CONSTANT MAGNITUDE CIRCLE (M- CIRCLE)

$$M(j\omega) = \frac{G(j\omega)}{1 + G(j\omega)}$$

$$G(j\omega) = x + jy$$

$$|M(j\omega)| = \frac{\sqrt{x^2 + y^2}}{\sqrt{(1+x)^2 + y^2}}$$

$$M^2(1+x)^2 + M^2 y^2 = x^2 + y^2$$

If M = 1, then from the above equation we obtain X = -1/2.This is the equation of a straight line parallel to the Y-axis and passing through the point (-1/2, 0).

$$x^2(1-M^2) + (1-M^2)y^2 - 2M^2 x = M^2$$

Divide both the sides by $(1\text{-}M^2)$

$$x^2 + y^2 - 2\frac{M^2}{1-M^2}x = \frac{M^2}{1-M^2}$$

$\left(\frac{M^2}{1-M^2}\right)^2$ **add in both sides**

$$x^2 + y^2 - 2x\frac{M^2}{1-M^2} + \left(\frac{M^2}{1-M^2}\right)^2 = \frac{M^2}{1-M^2} + \left(\frac{M^2}{1-M^2}\right)^2$$

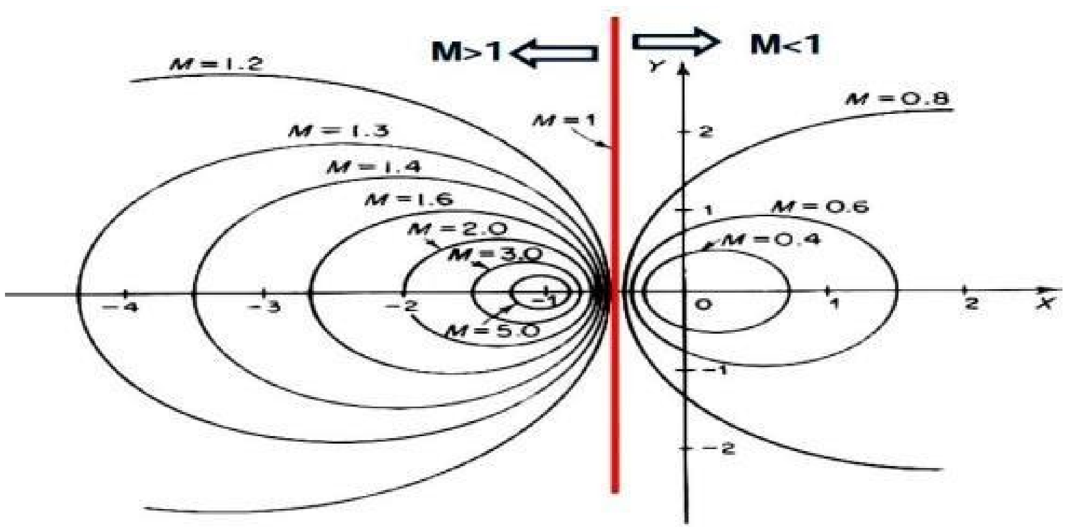
$$\left(x - \frac{M^2}{1-M^2}\right)^2 + (y-0)^2 = \frac{M^2}{1-M^2} + \frac{M^4}{(1-M^2)^2}$$

$$\left(x - \frac{M^2}{1-M^2}\right)^2 + (y-0)^2 = \frac{M^2}{(1-M^2)^2}$$

The above equation represents a family of circles with its center at $\left(\frac{M^2}{1-M^2}, 0\right)$ and radius $\left|\frac{M}{1-M^2}\right|$.

The constant M locii for different value of M. It is clear that:

i. The locii are symmetrical wrt to M= 1
ii. The M-circles for M>1 are on the left side of the line M=1 and for M<1 the constant M-circles are on the right side of the line M =1.



CONSTANT PHASE CIRCLE (N- CIRCLE):

$$\angle M = \alpha = \frac{\angle G(j\omega)}{\angle 1 + G(j\omega)}$$

$$\alpha = \tan^{-1}\frac{y}{x} - \tan^{-1}\frac{y}{1+x}$$

$$N = \tan\left(\tan^{-1}\frac{y}{x} - \tan^{-1}\frac{y}{1+x}\right)$$

$$\tan(A-B) = \frac{\tan A - \tan B}{1 + \tan A \tan B}$$

Here, $\tan(\alpha)=N$

$\therefore$ $$N = \frac{y}{x^2+x+y^2} \; ; \quad \text{or} \quad N(x^2+x+y^2) = y$$

$$x^2 + x + y^2 - \frac{y}{N} = 0$$

Add $\frac{1}{4} + \frac{1}{(2N)^2}$ **to both sides, we get**

$$x^2 + x + y^2 - \frac{y}{N} + \frac{1}{4} + \frac{1}{(2N)^2} = \frac{1}{4} + \frac{1}{(2N)^2}$$

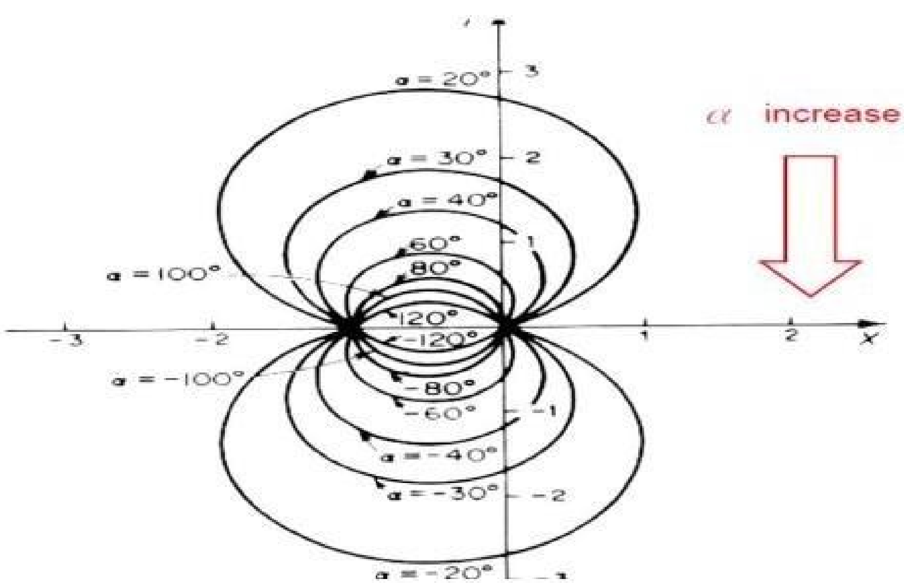
or $$(x+1/2)^2 + \left(y - \frac{1}{2N}\right)^2 = \frac{1}{4} + \frac{1}{(2N)^2}$$

The above equation represents a family of circles with its center at $(-\frac{1}{2}, \frac{1}{2N})$ and radius

$$\sqrt{\frac{1}{4} + \left(\frac{1}{2N}\right)^2}$$

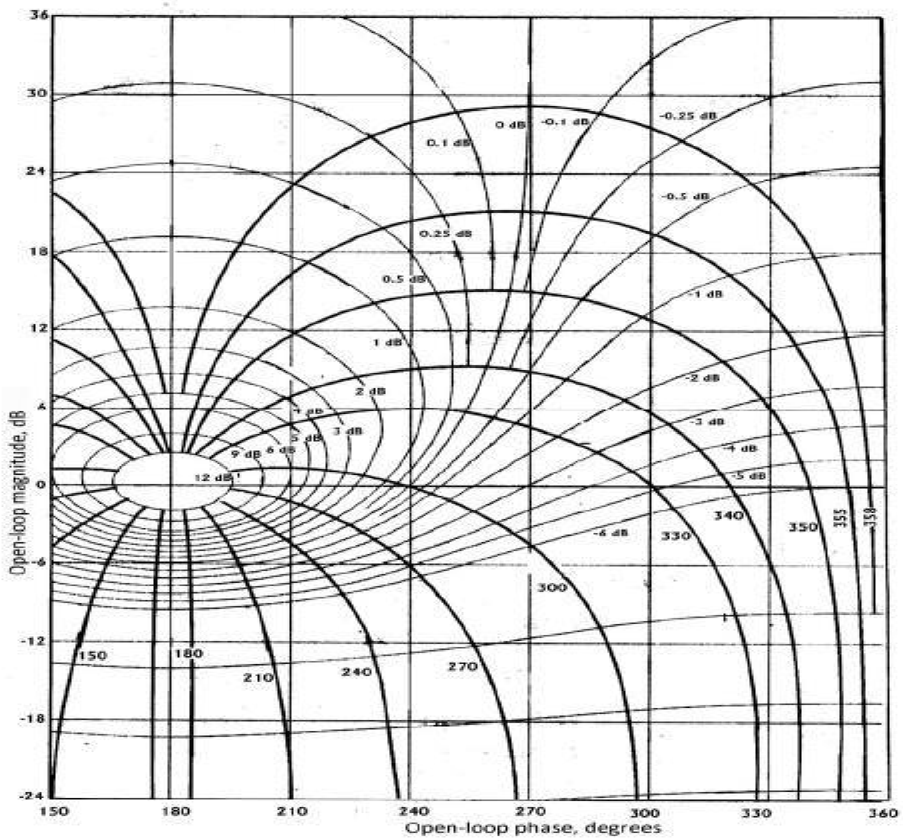
It is observed that

a) The centre is lying always at a distance x= -1/2 and y depends upon the phase shift.
b) All the circles passes through -1 as well as 0.



NICHOLS CHART

- The chart consisting of constant-magnitude loci and constant phase-angle loci in the log-magnitude versus phase diagram is called Nichols chart.
- The critical point (-1+j0) is mapped to the Nichols chart as the point (0 dB, 180degree). The Nichols chart contains curves of constant closed-loop magnitude and phase angle.
- The designer can graphically determine the phase margin, gain margin, resonant peak magnitude, resonant peak frequency, and bandwidth of the closed loops system from the plot of the open-loop locus.
- The Nichols chart is symmetric about -180 degree axis. The constant-magnitude loci and constant phase-angle loci repeat for every 360 degree, and there is a symmetry at every 180 degree. The constant-magnitude loci are centred about the critical point (0 dB, -180 degree).
- The intersection of the open-loop frequency response curve and the constant-magnitude loci and constant phase-angle loci give the values of the magnitude and the phase angle of the closed loop frequency response at each frequency point.
- If the open-loop frequency response curve does not intersect the constant-magnitude loci but is tangent to it, then the resonant peak value of the closed-loop frequency response is given by that loci. The resonant peak frequency is given by the frequency at the point of tangency.
- The phase crossover point is the point where the open-loop locus intersects the -180 degree axis, and the gain crossover point is the point where the locus intersects the 0 dB axis.
- The phase margin is the distance (measured in degrees) between the gain crossover point and the critical point (0 dB, -180 degrees).
- The gain margin is the distance (in decibels) between the phase crossover point and the critical point. The frequency at the intersection of the open-loop locus and the -3 dB locus gives the bandwidth.



36
30
24
18
12
6
0
-6
-12
-18
-24
150
180
210
240
270
300
330
360
Open-loop phase, degrees
Open-loop magnitude, dB
0 dB
-0.1 dB
-0.25 dB
0.1 dB
-0.5 dB
0.25 dB
0.5 dB
-1 dB
1 dB
-2 dB
2 dB
3 dB
-3 dB
-4 dB
-5 dB
9 dB
6 dB
5 dB
12 dB
-6 dB
340
350
355
358
300