

RGPVNOTES.IN

Program : **B.Tech**

Subject Name: **Basic Electrical & Electronics Engineering**

Subject Code: **BT-104**

Semester: **1st**



LIKE & FOLLOW US ON FACEBOOK
facebook.com/rgpvnotes.in

UNITS AND DIMENSIONS :

The quantitative description of any object or material involves measurement and comparison of physical quantities. Any physical quantity can be measured using a standard unit of that quantity. The unit of a physical quantity is the reference standard used to measure it.

Dimension is a measurable physical quantity, while unit is a way to assign a number or measurement to that dimension. There is difference between dimension and unit. For example, length is a dimension, but it is measured in units of feet or meter. A particular quantity can be reported in many different kinds of units, but it will always have the same dimensions. Dimensions are represented using symbols by: length [L], mass [M] and time [T]. In order to maintain uniformity in the field of science and engineering the S.I. unit is used ("ysteme International d'Unites). The seven fundamental units in "I are - Meter, Kilogram, Second, Ampere, Kelvin, Mole and Candela.

IMPORTANT LAWS :

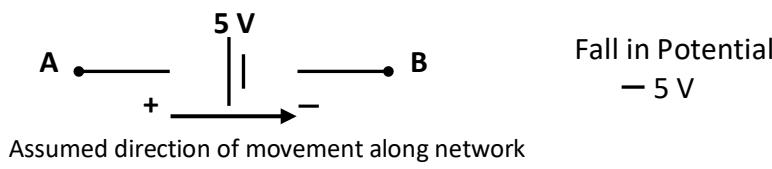
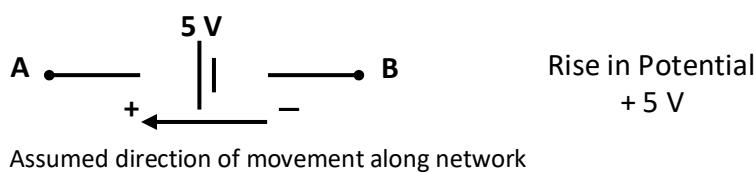
- a. **Ohm's law :** The current through a conductor between two points is directly proportional to the potential difference across it, provided the temperature of conductor and all other factors remain constant. The relations for ohms law are $I = \frac{V}{R}$; $V = IR$; $R = \frac{V}{I}$

- b. **Kirchhoff's current law :** The algebraic sum of the currents at a junction is equal to zero.

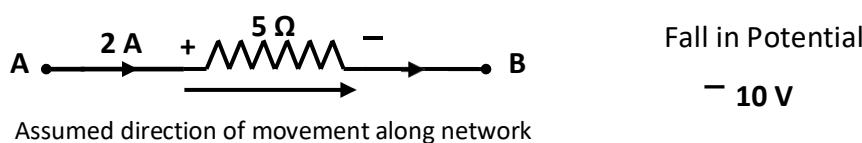
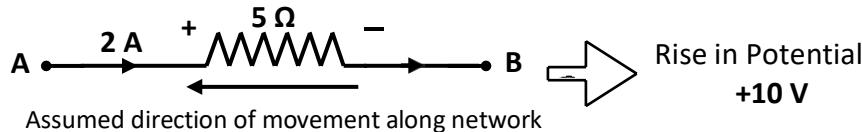
- c. **Kirchhoff's voltage law :** The algebraic sum of the voltage sources in any closed circuit is always equal to the sum of the voltage drops as well as voltage rises in that closed circuit.

- d. **Assumption of polarity (sign) while applying Kirchhoff's voltage law (KVL) to electrical networks :**

i) **Voltage sources :**



ii) **Passive elements:**



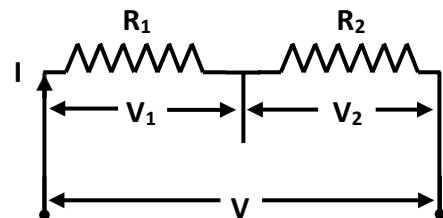
The analysis of DC circuits can be carried out if the below mentioned relations are known :

a. Resistors in series : $R_{eq} = R_1 + R_2$

b. Resistors in parallel : $R_{eq} = \frac{R_1 R_2}{R_1 + R_2}$

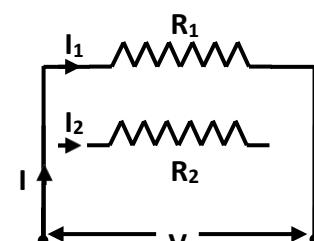
c. The voltage division for the circuit shown is :

$$V_1 = V \frac{R_1}{R_1 + R_2} \quad \text{and} \quad V_2 = V \frac{R_2}{R_1 + R_2}$$



d. The division of current for the circuit shown is :

$$I_1 = I \frac{R_2}{R_1 + R_2} \quad \text{and} \quad I_2 = I \frac{R_1}{R_1 + R_2}$$



In DC circuit analysis usually the circuits are reduced in steps to get their equivalent resistance and then obtain the required solution.

RELATIONS FOR POWER AND ENERGY:

We have electrical power, $P = VI$

An electrical power of 1 watt is consumed in a circuit if a potential difference of 1 volt when applied across it, causes a current of 1 ampere to flow through it. Other relations for power are, $P = I^2R$ and $P = V^2/R$.

Unit for power is watts or kW.

We have electrical energy, $E = \text{Power} \times \text{Time}$.

An electrical energy of 1 watt-sec is consumed in a circuit when a power of 1 watt is utilized for one second

OR An electrical energy of 1 kWh is consumed in a circuit when a power of 1 kW is utilized for one hour.

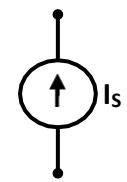
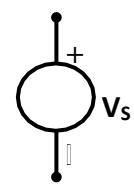
Other relations for energy are, $E = I^2Rt$ and $E = Vit$; Unit for energy is Watt-sec or Watt-hr or kWh.

VOLTAGE AND CURRENT SOURCES :

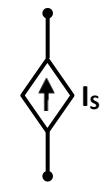
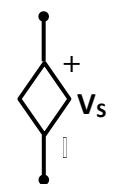
- Any device that produces electrical energy can be called a source.
- A source is usually expected to deliver power to a network and not to absorb it.
- A voltage source maintains the required difference in potential across the circuit it is connected.
- A current source supplies the required quantity of current to the circuit it is connected.
- An ideal constant voltage source is one whose output voltage remains absolutely constant irrespective of the change in load current. These voltage sources must possess zero internal resistance, so that the internal voltage drop in the source is zero. It is not practically possible to have an ideal constant voltage source.
- An ideal constant current source is one whose output current remains absolutely constant. These current sources have infinite internal resistance. Practically these sources possess a very high resistance when compared to its external load resistance.

DEPENDENT AND INDEPENDENT SOURCES :

- The sources in which the voltage or current depends upon a current or voltage elsewhere in the circuit are known as Dependent sources or Controlled sources.
- The sources in which the voltage is completely independent of the current or the current is completely independent of the voltage are known as Independent sources.
- An ideal independent voltage source is one that maintains a specified voltage between its source terminals regardless of the current drawn from it. *It is symbolised as shown :*
The positive and negative signs indicate the conventional direction of electric field when the source is applied to a load.
- An ideal independent current source is one that maintains a specified current through its terminals regardless of the voltage across the terminals. *It is symbolised as shown :*
The arrow indicates the conventional direction of current when the source is connected to a load.

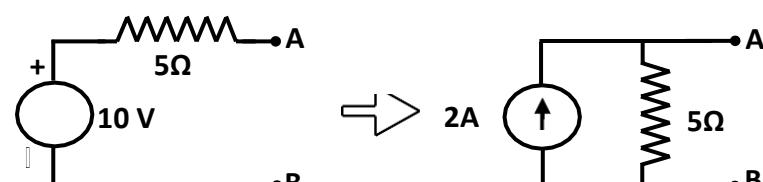


- A dependent voltage source is one that produces a voltage as a function of voltages elsewhere in a given circuit. *It is symbolised as shown :*
- A dependent current source is one that produces a current as a function of currents elsewhere in a given circuit. *It is symbolised as shown :*

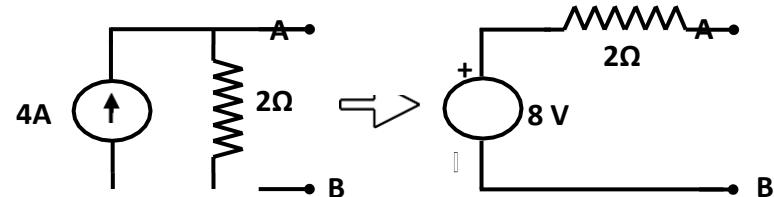


SOURCE CONVERSION :

- A voltage source with a series resistor can be converted into an equivalent current source with a resistor in parallel to it.
- A current source with a parallel resistor can be converted into an equivalent voltage source with a resistor in series with it.
- The conversions are possible only when their respective open circuit voltages are equal or their respective short circuit currents are equal.
- *Example for voltage source to current source conversion :*

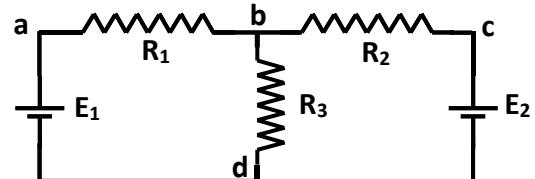


- *Example for current source to voltage source conversion :*
- Conversion of sources helps in simplifying the analysis of circuits.



NETWORK TERMINOLOGIES :

- A network or circuit is an arrangement of active and passive elements that form closed paths.
- Consider the circuit shown :
- It has two active elements E_1 and E_2 .
- It has three passive elements R_1 , R_2 and R_3 .
- A node of a network is an equi-potential surface at which two or more circuit elements are joined.
- In the circuit shown above a , b , c and d are nodes.
- A junction is that point in a network where three or more circuit elements are joined.
- In the circuit there are two junctions b and d .
- A branch is that part of a network which lies between junction points.
- There are three branches dab , dcb and db .
- The branch dab has two elements E_1 and R_1 .
- The branch dcb has two elements E_2 and R_2 .
- The third branch db has only one element R_3 .
- A loop is any closed path of a network.
- The loops in the circuit are $abda$, $dbcd$ and $abcd$.
- A mesh is the most elementary form of a loop.
- The meshes in the circuit are $abda$ and $dbcd$.
- A mesh is also a loop that cannot have another loop within it.
- A mesh current is that current which flows around the perimeter of the mesh.
- The mesh currents are always assumed to flow in the clockwise direction.
- Branch currents have physical identity but mesh currents are fictitious quantities introduced so that they allow us to solve problems with minimum number of unknowns.



MESH AND LOOP ANALYSIS :

- When the number of branches in a network increase, the earlier methods used will lead to complications. In order to simplify the solution of such networks one of the methods is the Loop analysis or the Mesh analysis.
- The step by step procedure adopted to use the method of mesh analysis is :
 - a. Observe the circuit for finding the possible number of meshes, if there are any current sources, convert them into their equivalent voltage sources.
 - b. Assign mesh currents to each mesh assuming the current to flow in clockwise direction.
 - c. Apply KVL to each mesh and write the equations.
 - d. The number of equations will be equal to the number of unknown mesh currents.

- e. The equations are solved to determine the mesh currents.
- f. The required branch currents are determined from the mesh currents determined.
- g. In case the branch current determined is negative, then the branch current is flowing opposite to the assumed direction.
- h. In case the branch current determined is positive, then the branch current is flowing in the assumed direction.

NODE VOLTAGE ANALYSIS :

- A node is a point in a network that is common to two or more circuit elements. If three or more elements are joined at a point, that point can be called a junction. It is also called as an independent node or principle node.
- Usually the negative terminal of an active element is selected as the reference node or datum node and its potential is assumed to be zero.
- The node voltage is the voltage of a given node with respect to the reference node or datum node.
- The node analysis method helps us to find the voltages at all the principle nodes with respect to the reference node.
- Usually all the branch currents are assumed to be positive when the direction of the currents are not known or not given in the circuit.
- At all the principle nodes, the currents flowing towards the node are considered negative and the currents flowing away from the node are considered positive.
- The step by step procedure adopted to use the method of nodal analysis is:
 - a. Observe the circuit to find the number of principle nodes and identify the reference node.
 - b. Number the principle nodes serially and assume the node voltages.
 - c. Assume the currents to flow outward from the nodes in each branch.
 - d. Apply KCL to all the nodes and write the equations in terms of voltages and resistances.
 - e. The number of equations will be equal to the number of principle nodes.
 - f. The equations are solved to find the values of the assumed node voltages.
 - g. With the determined values of the node voltages all the branch currents are calculated.

SUPERPOSITION THEOREM :

This theorem is very useful as it extends the use of Ohm's law to circuits that have more than one source. It is possible to calculate the effect of each source at a time and then superimpose results of all the other sources.

Statement : "In a network with two or more sources, the current or voltage for any component is the algebraic sum of the effects produced by each source acting separately".

Step by step procedure to analyse a network using superposition theorem :

Let us consider the circuit shown in figure-A, I_1 , I_2 and I_3 are the currents flowing in the circuit due to the two voltage sources of 8V and 12V.

To solve the circuit by using superposition theorem, only one voltage source has to be considered to be acting at a time in the circuit.

So the 8V source is retained and the 12V source is removed, as it has no internal resistance the circuit is drawn as shown in figure-B.

I_1' , I_2' and I_3' are the currents flowing in the circuit as shown due to the 8V source only. The equivalent resistance of the circuit is calculated, the total current and the branch currents are found using Ohm's law. Next considering the 12V source only in the circuit the 8V source is removed, as it has no internal resistance the circuit is drawn as shown in figure-C.

I_1'' , I_2'' and I_3'' are the currents flowing in the circuit as shown due to the 12 V source only. The total current and branch currents are calculated.

The currents I_1 , I_2 and I_3 flowing in the circuit shown in figure-A, can be obtained by combining the values of the currents flowing in figure-B and figure-C.

So the branch currents of figure-A are $I_1 = I_1' - I_1''$; $I_2 = I_2'' - I_2'$ and $I_3 = I_3' + I_3''$

THEVENIN'S THEOREM :

This theorem is quite useful in analyzing complicated networks comprising of a number of voltage or current sources. It helps in simplifying the process of solving for the unknown values of voltage and current in a network.

By Thevenin's theorem, many sources and components, no matter how they are interconnected, can be represented by an equivalent series circuit with respect to any pair of terminals in the network.

In fig. - 1 below the block at the left contains a network connected to terminals A and B, which can be replaced by a single source of emf, V_{TH} in series with a single resistance R_{TH} .

Where V_{TH} is the open circuit voltage across terminals A and B and R_{TH} is the open circuit resistance across terminals A and B

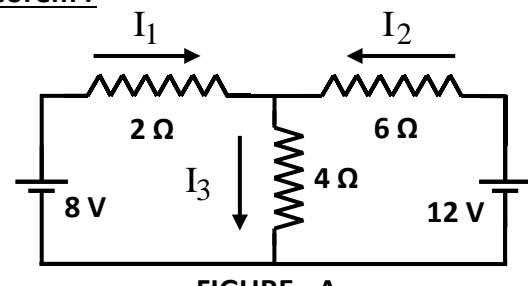


FIGURE - A

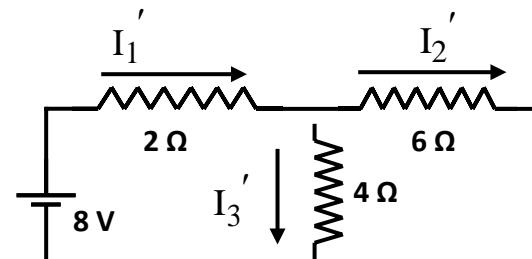


FIGURE - B

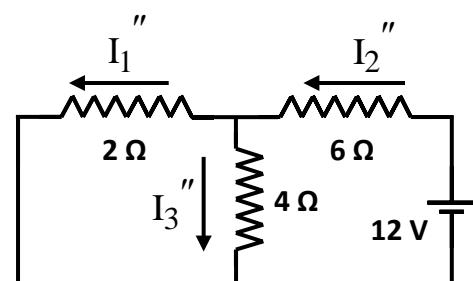


FIGURE - C

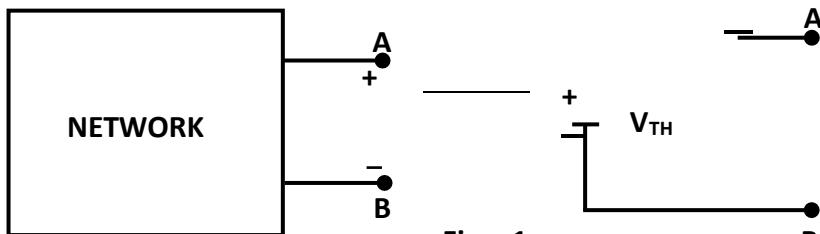


Fig. - 1

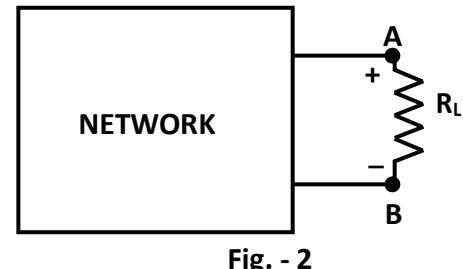


Fig. - 2

STATEMENT : “Thevenin theorem states that the entire network connected to A and B can be replaced by a single voltage source V_{TH} in series with a single resistance R_{TH} , connected to the same two terminals”.

STEP BY STEP PROCEDURE IN THEVENIZING A CIRCUIT :

The step by step procedure adopted to solve any network by Thevenin’s theorem is given below :

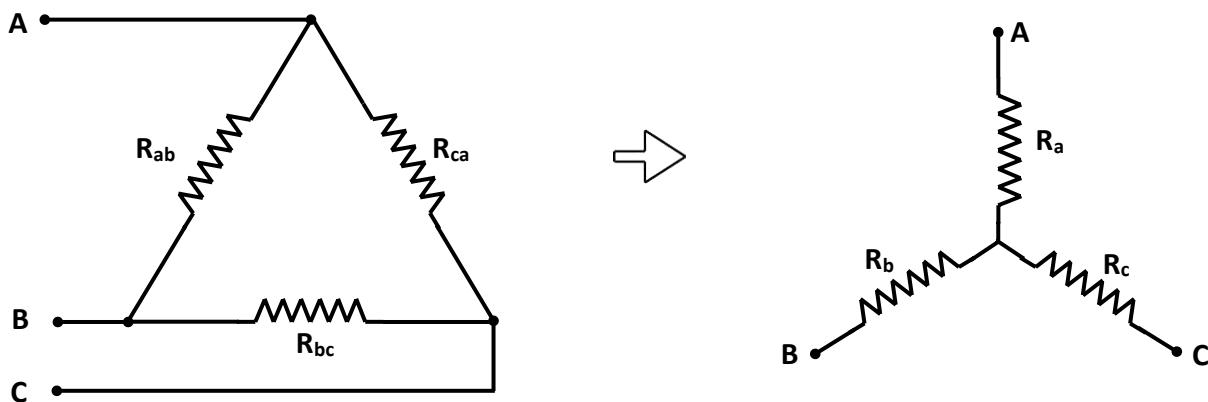
- The load resistor R_L of the network through which the current flowing has to be determined is identified.
- The load resistor R_L is temporarily disconnected from the network.
- Let the points be named A and B.
- The open circuit voltage which appears across the points A and B is determined. This is called Thevenin voltage V_{TH} .
- In order to determine the Thevenin resistance of the network behind the points A and B.
- The voltage sources in the network are replaced by their internal resistances and the current sources are replaced by an open circuit.
- The equivalent resistance across the terminals A and B is determined which is called Thevenin resistance R_{TH} .
- Replace the entire network by the Thevenin source, whose voltage is V_{TH} and whose internal resistance is the Thevenin resistance R_{TH} .
- Connect the load resistor R_L back across the points A and B, from where it was removed earlier.
- Calculate the current flowing through the load resistor using the relation :

$$I = \left(\frac{V_{TH}}{R_{TH} + R_L} \right)$$

STAR-DELTA RELATIONS :

- ❖ When a three terminal circuit is encountered in any network the star delta relations can be used to simplify the circuit.
- ❖ By initially converting the three terminal network from one form to another and by applying other simplifying techniques the network can be solved.

DELTA TO STAR CONVERSION :



Let us consider a Delta circuit shown above. Let us find the resistance between the terminals A and C with terminal B open. It is observed that the resistors R_{ab} and R_{bc} will be in series with each other and this series combination will be in parallel with R_{ca} . Hence, the equivalent resistance between the terminals A and C can be given as

$$\text{be written as, } R_{ca} = \frac{R_{ab} + R_{bc} + R_{ca}}{R_{ab}(R_{ca} + R_{bc})}$$

Similarly, resistance between terminals A and B, $R_{ab} = \frac{R_{ab} + R_{bc} + R_{ca}}{3}$

the resistance between terminals B and C, $R_{bc} = \frac{R_{bc}(R_{ab} + R_{ca})}{R_{ab} + R_{bc} + R_{ca}}$

Considering the Star circuit shown above, the resistance between the terminals A and C = $R_a + R_c$

Between A and B = $R_a + R_b$ and between B and C = $R_b + R_c$

Equating resistance between similar terminals in the two circuits, we get,

$$R_{cq} (R_{qb} + R_{bc})$$

$$R_a + R_b = \frac{R_{ab} (R_{ca} + R_{bc})}{R_{ab} + R_{bc} + R_{ca}} \dots \dots \dots (2)$$

$$R_b + R_c = \frac{R_{bc} (R_{ab} + R_{ca})}{R_{ab} + R_{bc} + R_{ca}} \dots \dots \dots (3)$$

$$R_{ab} \cdot R_{ca} - R_{ab} \cdot R_{bc}$$

Subtracting equation (3) from equation (1), we get, $R_a - R_b = \frac{R_{ab} + R_{bc} + R_{ca}}{R_{ab} + R_{bc} + R_{ca}}$ (4)

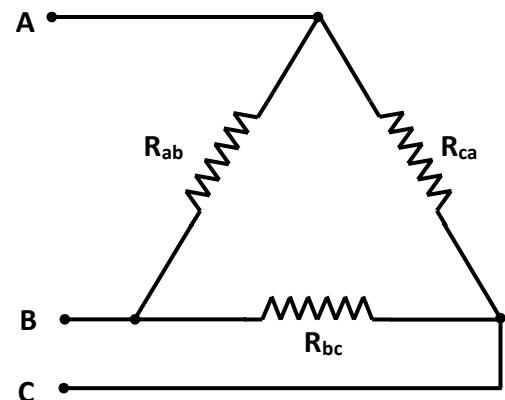
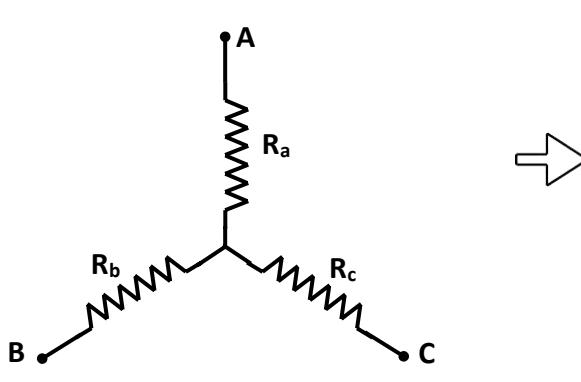
Adding equation (2) to equation (4) and dividing by 2, we get, $R_a = \frac{R_{ab} \cdot R_{ca}}{R_{ab} + R_{bc} + R_{ca}}$ (5)

Hence the Star values of resistors in terms of the Delta resistors are expressed as:

$$R_a = \frac{R_{ab} R_{ca}}{R_{ab} + R_{bc} + R_{ca}} ; \quad R_b = \frac{R_{bc} R_{ab}}{R_{ab} + R_{bc} + R_{ca}} ; \quad R_c = \frac{R_{ca} R_{bc}}{R_{ab} + R_{bc} + R_{ca}}$$

STAR TO DELTA CONVERSION :

Let us consider a Star connected circuit shown above. The resistance between the terminals A and C is found to be $R_a + R_c$. The resistance between B and C $= R_b + R_c$ and that between A and B $= R_a + R_b$.



Let us consider the Delta connected circuit shown above, the resistance between the terminals A and C with terminal B open can be written as, $R_{ac} = \frac{R_{ab} + R_{bc} + R_{ca}}{R_{ab}(R_{ca} + R_{bc})}$

$$\text{Similarly, resistance between terminals A and B, } R_{ab} = \frac{R_{ab}(R_{ca} + R_{bc})}{R_{ab} + R_{bc} + R_{ca}}$$

$$\text{and the resistance between terminals B and C, } R_{bc} = \frac{R_{bc}(R_{ab} + R_{ca})}{R_{ab} + R_{bc} + R_{ca}}$$

Equating resistance between similar terminals in the two circuits, we get,
 $R_{ca}(R_{ab} + R_{bc})$

$$R_a + R_c = \frac{R_{ab} + R_{bc} + R_{ca}}{R_{ab} + R_{bc} + R_{ca}} \dots\dots\dots(1)$$

$$R_a + R_b = \frac{R_{ab}(R_{ca} + R_{bc})}{R_{ab} + R_{bc} + R_{ca}} \dots\dots\dots(2)$$

$$R_b + R_c = \frac{R_{bc}(R_{ab} + R_{ca})}{R_{ab} + R_{bc} + R_{ca}} \dots\dots\dots(3)$$

$$R_{ab} \cdot R_{ca} - R_{ab} \cdot R_{bc}$$

$$\text{Subtracting equation (3) from equation (1), we get, } R_a - R_b = \frac{R_{ab} \cdot R_{ca} - R_{ab} \cdot R_{bc}}{R_{ab} + R_{bc} + R_{ca}} \dots\dots\dots(4)$$

$$\text{Adding equation (2) to equation (4) and dividing by 2, we get, } R_a = \frac{R_{ab} \cdot R_{ca}}{R_{ab} + R_{bc} + R_{ca}} \dots\dots\dots(5)$$

$$\text{Similarly, } R_b = \frac{R_{bc} R_{ab}}{R_{ab} + R_{bc} + R_{ca}}$$

$$R_c = \frac{R_{ca} R_{bc}}{R_{ab} + R_{bc} + R_{ca}}$$

From equations (5), (6) and (7), we get,

$$\begin{aligned}
 R_a R_b + R_b R_c + R_c R_a &= \frac{\left(R_{ab}^2 \cdot R_{bc} \cdot R_{ca} + R_{bc}^2 \cdot R_{ab} \cdot R_{ca} + R_{ca}^2 \cdot R_{ab} \cdot R_{bc} \right)}{(R_{ab} + R_{bc} + R_{ca})^2} \\
 R_a R_b + R_b R_c + R_c R_a &= \frac{(R_{ab} + R_{bc} + R_{ca})(R_{ab} \cdot R_{bc} \cdot R_{ca})}{(R_{ab} + R_{bc} + R_{ca})^2} \\
 R_a R_b + R_b R_c + R_c R_a &= \frac{(R_{ab} + R_{bc} + R_{ca})^2}{(R_{ab} \cdot R_{bc} \cdot R_{ca})} \quad(8)
 \end{aligned}$$

Dividing Equation (8) by equation (7), we get,

$$R_{ab} = \frac{(R_a R_b + R_b R_c + R_c R_a)}{R_c}$$

Dividing Equation (8) by equation (5), we get,

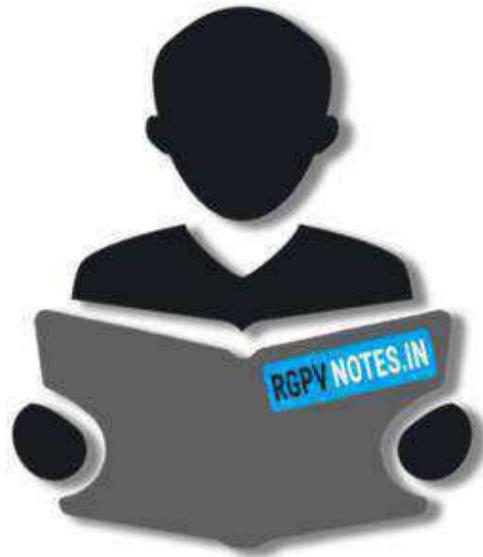
$$R_{bc} = \frac{(R_a R_b + R_b R_c + R_c R_a)}{R_a}$$

Dividing Equation (8) by equation (6), we get,

$$R_{ca} = \frac{(R_a R_b + R_b R_c + R_c R_a)}{R_b}$$

Hence the Delta values of resistors in terms of the Star resistors are expressed as:

$$R_{ab} = \frac{R_a R_b + R_b R_c + R_c R_a}{R_c}; R_{bc} = \frac{R_a R_b + R_b R_c + R_c R_a}{R_a}; R_{ca} = \frac{R_a R_b + R_b R_c + R_c R_a}{R_b}$$



RGPVNOTES.IN

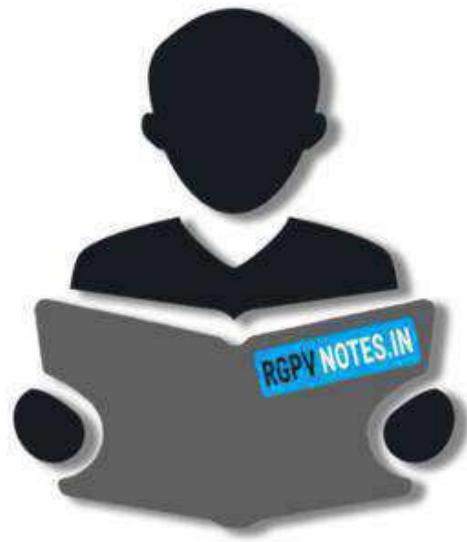
We hope you find these notes useful.

You can get previous year question papers at
<https://qp.rgpvnotes.in> .

If you have any queries or you want to submit your
study notes please write us at
rgpvnotes.in@gmail.com



LIKE & FOLLOW US ON FACEBOOK
facebook.com/rgpvnotes.in



RGPVNOTES.IN

Program : **B.Tech**

Subject Name: **Engineering Chemistry**

Subject Code: **BT-101**

Semester: **1st**



LIKE & FOLLOW US ON FACEBOOK
facebook.com/rgpvnotes.in

ENGINEERING CHEMISTRY (BT-101) NOTES

UNIT I: WATER ANALYSIS, TREATMENTS AND INDUSTRIAL APPLICATIONS

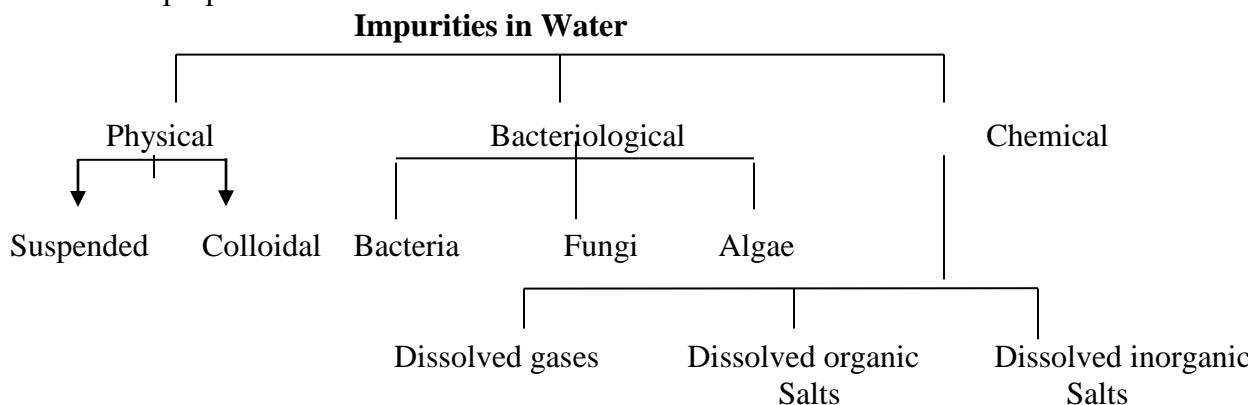
INTRODUCTION: For the existence of all living beings, water is very crucial. Almost all human activities – domestic, agricultural and industrial demand use of water. Although water is nature's most wonderful and abundant compound but only less than 1% of the world's water resources is available for ready use. Hence, water has to be used carefully and economically.

SPECIFICATIONS OF WATER:

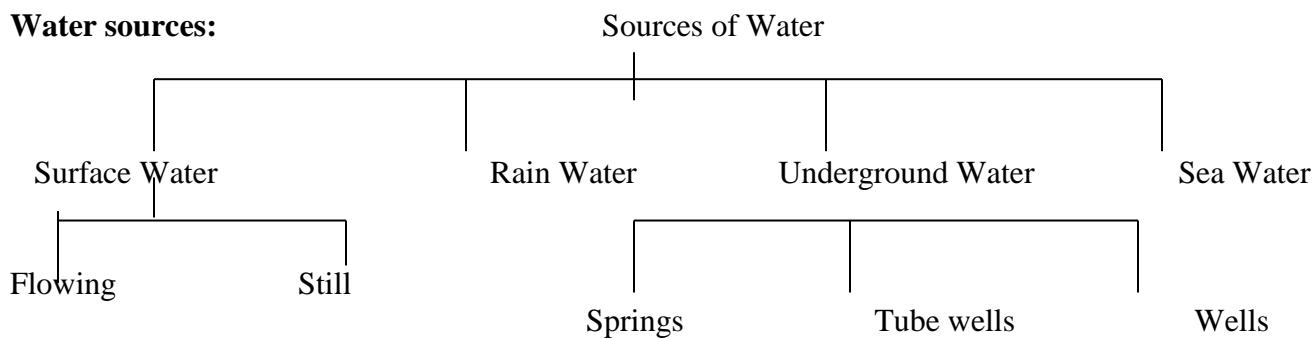
Different uses of water demand different specifications –

- (i) Textile industry needs frequent dying of clothes and the water used by this industry should be soft and free from organic matter. Hard water decreases the solubility of acidic dyes. Organic matter imparts foal smell.
- (ii) Laundries require soft water, free from colour, Mn and Fe, because hardness inc. consumption of soaps, salts of Fe and Mn impart a grey or yellow shade to the fabric.
- (iii) Boilers require water of zero hardness otherwise efficient heat transfers is prevented by scale formation. Untreated water can lead to corrosion of boiler material.
- (iv) Paper industry requires water free from SiO_2 as it produces cracks in paper; turbidity as it can affect brightness and colour of paper; alkalinity as it consumes more alum; hardness as Ca^{2+} Mg^{2+} . Salts increases the ash content of the paper.
- (v) Sugar industry requires water free from hardness because hard water causes difficulty in the crystallization of sugar.
- (vi) Dairies and pharmaceutical industry require ultra pure water, which should be colorless, tasteless, odorless and free from pathogenic organisms.

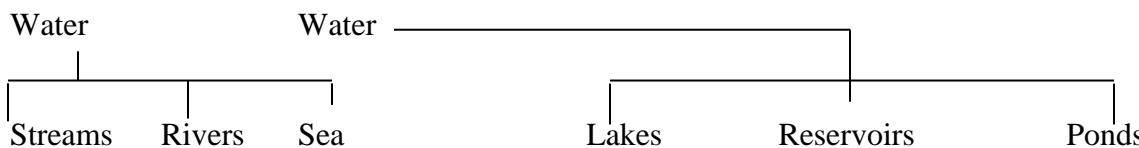
Therefore water needs to be treated to remove undesirable impurities. "Water treatment" is the process by which all types of undesirable impurities are removed from water and making it fit for domestic or industrial purposes.



Water sources:



ENGINEERING CHEMISTRY (BT-101) NOTES

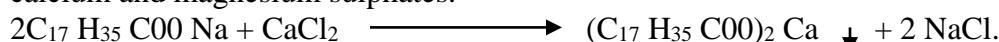


River water contains dissolved minerals like chlorides sulphates, bicarbonates of sodium, magnesium, calcium and iron. Its composition is not constant. Lake water has high quantity of organic matter present in it. Its chemical composition is also constant. Rain water, in the purest form of natural water. When it comes down, it dissolves organic and inorganic suspended particles and some amount of industrial gases.

Underground water is free from organic impurities and is clearer in appearance due to filtering action of the soil. It has large amount of dissolved salts. Sea water is very impure due to continuous evaporation and impurity thrown by rivers as they join sea.

HARDNESS OF WATER:

Hardness is defined as soap consuming capacity of water sample. It is that characteristic "which prevent the lathering of soap." It is due to presence of certain salts of Ca, Mg and other heavy metal ions like Al^{3+} , Fe^{3+} and Mn^{2+} dissolved in it. A sample of hard water, when treated with soap (K or Na salt of higher fatty acids like oleic, palmitic or stearic acid), does not produce lather, but forms insol. white scum or ppt. which does not possess any detergent action, due to formation of insoluble soaps of calcium and magnesium sulphates.



Soap (Hardness) Calcium Sterate (Insoluble)



Soap (Hardness) Magnesium Sterate (Insoluble)

S.No	HARD WATER	SOFT WATER
1.	Water which does not produce lather with soap solution readily, but forms a ppt.	Water which lather easily on shaking with soap solution, is called soft water.
2.	It contains dissolved Ca & Mg salts in it.	It does not contain Ca & Mg salts in it.
3.	Cleansing quality is depressed and lot of soap is wasted.	Cleansing quality is not depressed and so not soap is wasted.
4.	Boiling point of water is elevated, and more fuel and time are required for cooking.	Less fuel and time are required for cooking in soft water.
5.	Water is said to hard when hardness is above 100 mg. / ltr.	In soft water hardness is below 100 mg. / ltr.

TYPES OF HARDNESS: It is of following types

1. Temporary Hardness:

- (a) It is caused by presence of dissolved bicarbonates of Ca, Mg and other heavy metals and the carbonates of Iron. Example – $\text{Ca}(\text{HCO}_3)_2$ and $\text{Mg}(\text{HCO}_3)_2$.

- (b) It can be removed by boiling of water, when bicarbonates decompose to yield insoluble carbonates or hydroxides, which gets deposited as a crust at the bottom of vessel.



- (c) It is also known as carbonate hardness or alkaline hardness.

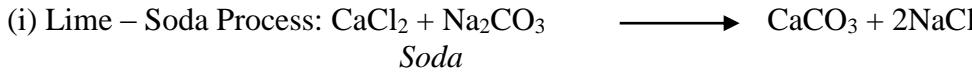
- (d) It is determined by titration with HCl using methyl orange as indicator.

ENGINEERING CHEMISTRY (BT-101) NOTES

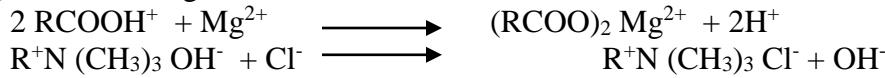
2. Permanent Hardness:

(a) It is due to presence of dissolved chlorides and sulphates of calcium, magnesium, iron and other heavy metals, eg. CaCl_2 , MgCl_2 , CaSO_4 , MgSO_4 , FeSO_4 , $\text{Al}_2(\text{SO}_4)_3$ etc.

(b) It cannot be destroyed by boiling. It can be removed by-



(iii) Ion – exchange Process:



(c) Permanent hardness is also known as non-carbonate or non-alkaline hardness.

3. Total Hardness = Carbonate Hardness (Temporary) + Non-carbonate Hardness (Permanent)

DEGREE OF HARDNESS:

Although hardness of water is never present in form of calcium carbonate because it is insoluble in water and calcium is not the only cation causing hardness but hardness is expressed in terms of equivalent amount (equivalents) of CaCO_3 i.e., the mg CaCO_3 / Litre.

The CaCO_3 is chosen as standard because:

1. CaCO_3 is chosen as the standard mainly because it is having great convenience to calculate as the molecular weight is 100 and its equivalent weight is 50.
2. It is sparingly soluble salt in water, thus, can be easily precipitated in the lime soda processes of softening water.

Therefore, all the hardness-causing impurities are first converted in terms of their respective weights equivalent to CaCO_3 and is expressed in parts per million.

Multiplication factors for different salts are:

Constituent Salt / ion	Molar Mass	n – factor	Chemical equivalent = molar mass	Multiplication factor for converting into equivalents of CaCO_3
			n – factor	
$\text{Ca}(\text{HCO}_3)_2$	162	(divalent) 2	162/2 = 81	100/2 x 81 = 100/162
$\text{Mg}(\text{HCO}_3)_2$	146	2	146/2 = 73	100/2 x 73 = 100/146
CaSO_4	136	2	136/2 = 68	100/2 x 68 = 100/136
MgSO_4	120	2	120/2 = 60	100/2 x 60 = 100/120
CaCl_2	111	2	111/2 = 47.5	100/2 x 47.5 = 100/111
MgCl_2	95	2	95/2 = 47.5	100/2 x 50 = 100/95
CaCO_3	100	2	100/2 = 50	100/2 x 50 = 100/100
MgCO_3	84	2	84/2 = 42	100/2 x 42 = 100/84
CO_2	44	2	44/2 = 22	100/2 x 22 = 100/44
$\text{Mg}(\text{NO}_3)_2$	148	2	148/2 = 74	100/2 x 74 = 100/148
HCO_3	61	(monovalent) 1	61/1 = 61	100/2 x 61 = 100/122
OH^-	17	1	17/1 = 17	100/2 x 17 = 100/34
CO_3^{2-}	60	2	60/2 = 30	100/2 x 30 = 100/60
NaAlO_2	82	1	82/2 = 82	82/2 x 82 = 100/164

ENGINEERING CHEMISTRY (BT-101) NOTES

$\text{Al}_2(\text{SO}_4)_3$	342	6	$342/6 = 57$	$100/2 \times 57 = 100/114$
$\text{FeSO}_4 \cdot 7\text{H}_2\text{O}$	278	2	$278/2 = 139$	$100/2 \times 139 = 100/278$
H^+	1	1	$1/1 = 1$	$100/2 \times 1 = 100/2$

UNITS OF HARDNESS:

Parts per million (ppm): ppm is the parts of calcium carbonate equivalent hardness per 10^6 parts of water.

- Milligrams per litre (mg/L): It is the number of milligrams of CaCO_3 equivalent hardness present per litre of water.

$$1 \text{ mg / L.} = 1 \text{ mg of } \text{CaCO}_3 \text{ eq. Hardness / L of water}$$

But 1 L of water weights = 1000 gms.

$$= 1000 \times 1000 \text{ mg.}$$

$$1 \text{ mg / L} = 1 \text{ mg / } 10^6 \text{ mg} = 1 \text{ ppm.}$$

- Clarke's degree (${}^0\text{Cl}$): It is the number of grains of CaCO_3 equivalent hardness per gallon of water. It is the parts of CaCO_3 equivalent hardness per 70,000 parts of water.

- Degree French (${}^0\text{Fr}$): It is the parts of CaCO_3 eq. Hardness per 10^5 parts of water.

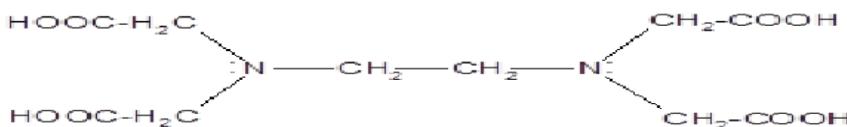
Relationship between units:

$$\begin{array}{lllll} 1 \text{ PPM} & = & 1 \text{ mg / L} & = & 0.1 {}^0\text{Fr} \\ 1 {}^0\text{Fr} & = & 10 \text{ PPM} & = & 10 \text{ mg / L} \\ 1 {}^0\text{Cl} & = & 14.3 \text{ PPM} & = & 14.3 \text{ mg/L} \end{array} = 0.07 {}^0\text{Cl} = 0.7 {}^0\text{Cl} = 1.43 {}^0\text{Fr}$$

Determination of Hardness by EDTA Method

- EDTA is abbreviation of Ethylene diamine tetra acetic acid.
- EDTA dissolves in water with great difficulty and in a very very small quantity.
- On the contrary its di-sodium salt dissolves in water quickly and completely. Hence for common experimental purpose, in place of EDTA, its di-sodium derivative is used.
- EDTA is a hexadentate ligand. It binds the metal ions in water i.e Ca^{2+} or Mg^{2+} to give highly stable chelate complex. (These metal ions are bonded via oxygen or nitrogen from EDTA molecule). Therefore this method is called as *Complexometric Titration*.

Structure of EDTA (Ethylene diamine Tetra acetic acid)

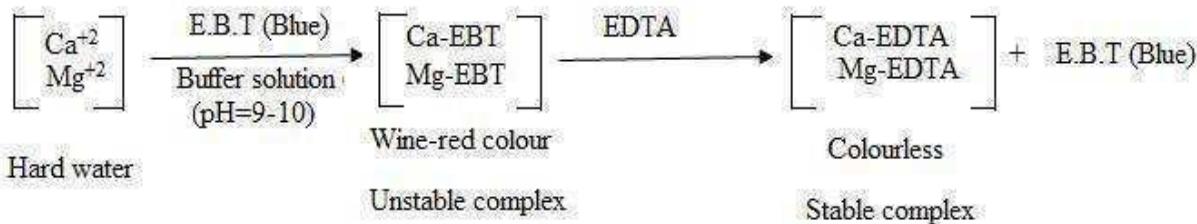


Principle of EDTA Method:

- The di-sodium salt of EDTA forms complexes with Ca^{2+} and Mg^{2+} , as well as with many other metal cations, in aqueous solution.

ENGINEERING CHEMISTRY (BT-101) NOTES

- Thus, the total hardness of a hard water sample, can be determined by titrating Ca^{2+} and Mg^{2+} present in the sample with di-sodium salt of EDTA (Na_2EDTA) solution, using ammonical buffer solution containing $\text{NH}_4\text{Cl}-\text{NH}_4\text{OH}$ of pH 10 using Eriochrome Black-T (EBT) as the metal indicator.
 - At pH 10, EBT indicator from wine red coloured unstable complex with Ca^{2+} and Mg^{2+} ions in hard water.
 - This complex is broken by EDTA solution during titration, giving stable complex with ions and releasing EBT indicator solution which is blue in colour. Hence the colour change is from wine red to blue (EBT's own colour).
 - Thus noting the colour change, the point of equivalence can be trapped and hardness of water can be determined by this method.



Advantages of EDTA Method:

- Greater accuracy
 - Highly rapid
 - Highly convenient

Alkalinity and its determination:

Alkalinity:

- It can be defined as “the concentration of the salts present in water which increases the concentration of OH^- ions due to hydrolysis thereby rising pH of water to alkaline range”.
 - Natural water when found alkaline, it is generally due to the presence of HCO_3^- , SiO_3^{2-} and sometimes CO_3^{2-} ions. In addition to the above the alkalinity of boiler water is also due to the presence of OH^- & PO_4^{2-} ions.
 - The extent of alkalinity depends on the presence of ions, which broadly can be categorized as presence of
 - (i) OH^- only (ii) CO_3^{2-} only (iii) HCO_3^- only (iv) OH^- & CO_3^{2-} together (v) HCO_3^- & CO_3^{2-} together.
 - Hydroxide and bicarbonates do not exist together because hydroxyl ions react with bicarbonate ions to form carbonate ions. Therefore existence of hydroxyl and bicarbonates ions together is ruled out.

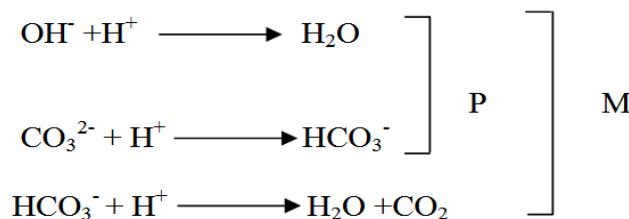


The alkalinity of natural water is due to the presence of hydroxides, carbonates and bicarbonates of Calcium and Magnesium. Alkalinity is a measure of the ability of water to neutralize the acids. The constituents causing alkalinity in natural water are as follows:

ENGINEERING CHEMISTRY (BT-101) NOTES

This is determined by titrating the sample with a standard solution of a strong acid. When the pH of the sample is above 8.3, titration is first carried out using phenolphthalein indicator. At the end point when the indicator changes from pink to colorless, the pH is lowered to about 4.5 due to addition of HCl. At this point complete neutralization of hydroxide and conversion of all the carbonate into bicarbonate occurs. The alkalinity measured up to this point is called phenolphthalein alkalinity. [P] Titration is continued using methyl orange indicator. The color changes from yellow to red and shows complete neutralization of all the bicarbonate ions.

When standard acid solution is added to alkaline water following reactions takes place:-



The total volume of acid used in both the stages corresponds to the neutralization of hydroxide, carbonate and bicarbonate and is thus, a measure of Total Alkalinity. [M]

Calculation of Alkalinity of water by following table

Alkalinity	OH^- (ppm)	CO_3^{2-} (ppm)	HCO_3^- (ppm)
$P = 0$	0	0	M
$P = 1/2M$	0	2P	0
$P < 1/2M$	0	2P	(M-2P)
$P > 1/2M$	(2P-M)	2(M-P)	0
$P = M$	P = M	0	0

Significance:

- For calculating the amounts of lime and soda required for water softening.
- In conditioning boiler feed water, highly alkaline waters may lead to Caustic Embrittlement and also may cause deposition of precipitates and sludge in boiler tubes and pipes.
- Bicarbonates of calcium and magnesium induce temporary hardness in water, which if untreated, causes scale formation in boilers.

Numerical based on Hardness & Strength

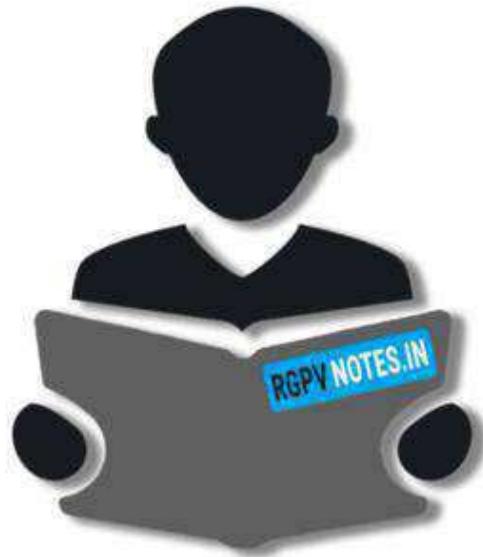
Formula for Determination of Hardness:

$$\text{Hardness} = \frac{\text{Strength (in mg/l)} \times \text{Chemical equivalent wt. of } \text{CaCO}_3}{\text{Chemical equivalent wt. of hardness producing salt}}$$

$$\text{Chemical equivalent wt. of hardness producing salt}$$

Formula for Determination of Strength:

$$\text{Strength} = \frac{\text{Hardness (in mg/l)} \times \text{Chemical equivalent wt. of hardness producing salt}}{\text{Chemical equivalent wt. of CaCO}_3}$$



RGPVNOTES.IN

We hope you find these notes useful.

You can get previous year question papers at
<https://qp.rgpvnotes.in> .

If you have any queries or you want to submit your
study notes please write us at
rgpvnotes.in@gmail.com



LIKE & FOLLOW US ON FACEBOOK
facebook.com/rgpvnotes.in



RGPVNOTES.IN

Program : **B.Tech**

Subject Name: **Basic Electrical & Electronics Engineering**

Subject Code: **BT-104**

Semester: **1st**



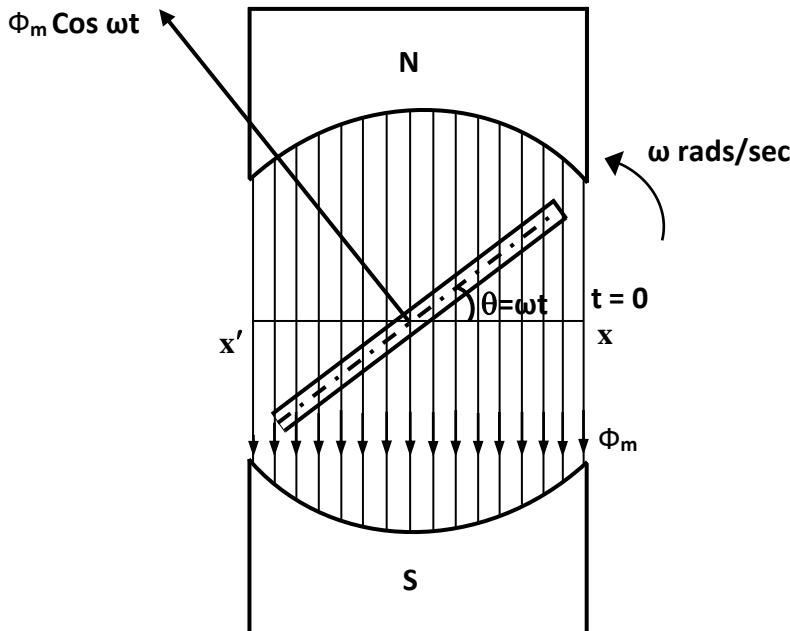
LIKE & FOLLOW US ON FACEBOOK
facebook.com/rgpvnotes.in

Unit - II Single phase AC Circuits - Syllabus

Generation of sinusoidal AC voltage, definition of average value, R.M.S. value, form factor and peak factor of AC quantity, Concept of Phasor, Concept of Power factor, Concept of impedance and admittance, Active, reactive and apparent power, analysis of R-L, R-C, R-L-C series and parallel circuit.

Single Phase AC Circuits

Generation of single phase voltages:



Let us consider a coil with N turns, rotating in a uniform magnetic field with an angular velocity ω rads/sec as shown in figure. Let the time be measured from the x -axis. When the plane of the coil coincides with the x -axis maximum flux Φ_m links the coil. After time t the coil moves through an angle $\theta = \omega t$

In this position the component of the flux which is perpendicular to the plane of the coil is $\Phi = \Phi_m \cos \omega t$

We have flux linkages $= N \Phi = N \Phi_m \cos \omega t$; Induced emf at this instant is, $e = -\frac{d}{dt} (N\Phi)$

$$= -N \frac{d}{dt} (\Phi_m \cos \omega t) = -\omega N \Phi_m (-\sin \omega t) = \omega N \Phi_m \sin \omega t$$

When the coil makes an angle $\theta = 90^\circ$, $\sin \theta = 1$, Hence, the emf induced in the coil is maximum, ie. E_m

$$\therefore E_m = \omega N \Phi_m \quad \text{or} \quad e = E_m \sin \omega t \quad \text{or} \quad e = E_m \sin \theta$$

The induced emf varies as sine function of the time angle ωt and when emf is plotted against time a sine curve is traced.

Terms and Definitions:

Cycle: One complete set of positive and negative values of an alternating quantity is known as a cycle.

Frequency: The number of cycles/sec is called the frequency of the alternating quantity.

Time period: The time taken by an alternating quantity to complete one cycle is called its time period.

Amplitude: The maximum value either positive or negative of an alternating quantity is known as its amplitude.

Phase: It is the fraction of the time period of that alternating quantity which has elapsed since the current last passed through the zero position of reference.

Phase difference: It is the difference in phase angle between any two alternating quantities.

Instantaneous value: It is the value of any alternating quantity at a particular instant of time.

Average value: It is that value of steady current, which transfers across any circuit the same charge as is transferred by that alternating current during the same time. Average value, $I_{av} = 0.637 I_m$

Root mean square value: It is given by that steady current which when flowing through a given circuit for a given time produces the same heat as produced by the alternating current when flowing through the same circuit for the same time. Root mean square value, $I_{rms} = 0.707 I_m$

Form factor: It is defined as the ratio of its rms value and average value

Form factor = rms value / average value = $0.707 I_m / 0.637 I_m = 1.11$ for sine wave.

Peak factor or Crest factor: It is defined as the ratio of its maximum value and rms value.

Peak factor = Maximum value / rms value = $I_m / 0.707 I_m = 1.414$ for sine wave.

Resistance: Its property is to oppose the flow of current through it. Resistance is a measurable quantity and its unit is Ohm.

Inductance: Its property is to induce emf in itself whenever a changing current flows through it, its unit is Henry.

Inductive reactance: It causes opposition to the flow of current through it. Reactance is a non-measurable quantity which can only be calculated its unit is Ohm.

Capacitance: It is the capacity of any capacitor to store charge and its unit is Farad.

Capacitive reactance: It causes opposition to the flow of current through it. Reactance is a non-measurable quantity which can only be calculated its unit is Ohm.

Impedance: It is the total opposition due to the resistance as well as the reactance of the circuit to the flow of current. It is a non-measurable quantity which can only be calculated its unit is Ohm.

Admittance: Admittance is a measure of how much current is admitted in a circuit. Admittance (Y) is the inverse of impedance (Z). Admittance has its most obvious utility in dealing with parallel AC circuits. The unit of admittance is Siemens.

Active power: It is also called Average power or True power or Real power. It is the actual power which is dissipated in the resistance of the circuit $P = VI \cos \Phi$ or $P = I^2R$, the unit is watts or KW.

Reactive power: It is the power developed in the inductive reactance of the circuit. It is also called wattless power Reactive power = $VI \sin \Phi$, its unit is Reactive Voltamperes or Kilovoltamperes reactive

Apparent power: It is the product of the rms values of voltage and current.

Apparent power = VI , its unit is Voltampere or Kilovoltamperes

Power factor: It is the cosine of the phase angle Φ existing between the voltage and current in any AC circuit. It has no unit. It can have values varying between zero and unity. It can also be either lagging or leading in nature.

Phasor: A phasor is a complex number representing a sinusoidal function whose amplitude, angular velocity, and initial phase are time invariant.

AC circuit with Resistance only:

Let the alternating voltage be $v = V_m \sin \omega t$

$$\text{the current } i = \frac{V_m}{R} \quad \text{or} \quad i = \frac{V_m \sin \omega t}{R}$$

The value of i will be maximum when $\sin \omega t = 1$

$$\therefore I_m = \frac{V_m}{R} \quad \text{or} \quad i = I_m \sin \omega t$$

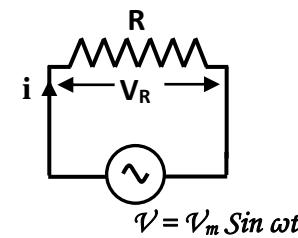
The voltage across the resistor and the current through the resistor are in phase with each other.

The instantaneous power, $p = vi = V_m \sin \omega t I_m \sin \omega t$

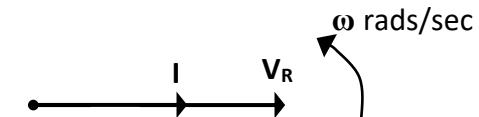
$$= V_m I_m \sin^2 \omega t = V_m I_m \left[\frac{1 - \cos 2\omega t}{2} \right] = \frac{V_m I_m}{2} - \frac{V_m I_m}{2} \cos 2\omega t$$

The average value of $\frac{V_m I_m}{2} \cos 2\omega t$ over a complete cycle is zero

$$\therefore P = \frac{V_m I_m}{2} \text{ or } = \frac{V_m}{\sqrt{2}} \frac{I_m}{\sqrt{2}} \therefore \text{Power} = VI \quad \text{Hence, a pure resistive circuit consumes power.}$$



CIRCUIT DIAGRAM



PHASOR DIAGRAM

AC circuit with Inductance only:

Let the alternating voltage be $v = V_m \sin \omega t$

$$\text{the self induced emf, } e_L = L \frac{di}{dt}$$

$$\therefore V_m \sin \omega t = L \frac{di}{dt} \quad \text{or} \quad di = \frac{V_m}{L} \sin \omega t dt$$

Integrating both sides, we have

$$i = \frac{V_m}{L} \int \sin \omega t dt = \frac{V_m}{\omega L} [-\cos \omega t] \quad \text{or} \quad i = \frac{V_m}{\omega L} \sin \left[\omega t - \frac{\pi}{2} \right]$$

We have i to be maximum when $\sin \left[\omega t - \frac{\pi}{2} \right]$ is unity

$$\therefore I_m = \frac{V_m}{\omega L} \quad \text{or} \quad i = I_m \sin \left[\omega t - \frac{\pi}{2} \right]$$

It is observed that the current lags the voltage by 90°

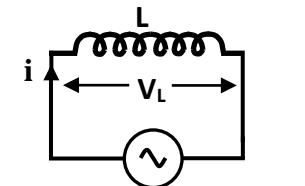
The quantity ωL is called inductive reactance and also represented as $X_L = 2\pi f L$

$$\text{The instantaneous power, } p = vi = V_m \sin \omega t I_m \sin \left[\omega t - \frac{\pi}{2} \right] = -V_m \sin \omega t I_m \cos \omega t$$

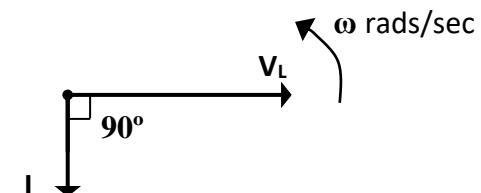
$$= -V_m I_m \sin \omega t \cos \omega t = -\frac{V_m I_m}{2} \sin 2\omega t$$

$$\text{Power for the complete cycle, } P = -\frac{V_m I_m}{2} \int_0^{2\pi} \sin 2\omega t dt = 0$$

A pure inductive circuit does not consume any power



CIRCUIT DIAGRAM



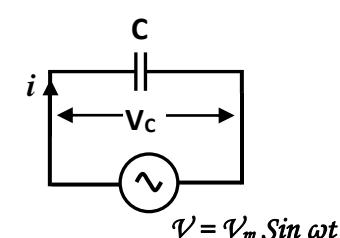
PHASOR DIAGRAM

AC circuit with Capacitance only:

Let the alternating voltage be $v = V_m \sin \omega t$

Let charge on the plates be $= q$

But charge, $q = Cv = CV_m \sin \omega t$



$V = V_m \sin \omega t$

Current, $i = \frac{dq}{dt} = \frac{d}{dt}[CV_m \sin \omega t] = \omega CV_m \cos \omega t$

or $i = \frac{V_m}{1/\omega C} \cos \omega t$ or $i = \frac{V_m}{1/\omega C} \sin \left[\omega t + \frac{\pi}{2} \right]$

The current i will be maximum when $\sin \left[\omega t + \frac{\pi}{2} \right]$ is unity

$\therefore I_m = \frac{V_m}{1/\omega C} = \frac{V_m}{X_C}$ Where X_C is the capacitive reactance

$\therefore i = I_m \sin \left[\omega t + \frac{\pi}{2} \right]$ It is observed that the current leads the voltage by 90°

The instantaneous power, $p = vi = V_m \sin \omega t I_m \sin \left[\omega t + \frac{\pi}{2} \right] = V_m \sin \omega t I_m \cos \omega t$

$= V_m I_m \sin \omega t \cos \omega t = \frac{V_m I_m}{2} \sin 2\omega t$

Power for the complete cycle, $P = \frac{V_m I_m}{2} \int_0^{2\pi} \sin 2\omega t dt = 0$

A pure capacitive circuit does not consume any power

R - L series circuit:

Let us consider a resistor and inductor in series.

If V is the rms value of the applied voltage then I

will be the rms value of the current drawn by the circuit.

The voltage across $R = V_R = IR$, where V_R is in phase with I .

The voltage across $L = V_L = IX_L$, where V_L leads I by 90°

The applied voltage V is the phasor sum of the two

voltage drops V_R and V_L

From the phasor diagram, we have

$$V = \sqrt{V_R^2 + V_L^2} = \sqrt{IR^2 + IX_L^2} = I \sqrt{R^2 + X_L^2}$$

$$I = \frac{V}{\sqrt{R^2 + X_L^2}} \quad \text{or} \quad I = \frac{V}{Z}$$

Where Z is the impedance of the circuit = $\sqrt{R^2 + X_L^2}$

$$\text{From the phasor diagram, } \tan \Phi = \frac{V_L}{V_R} = \frac{IX_L}{IR} = \frac{X_L}{R}$$

$$\text{or } \Phi = \tan^{-1} \frac{X_L}{R} \quad \text{Also } \cos \Phi = \frac{V_R}{V} = \frac{IR}{IZ} = \frac{R}{Z} \therefore \cos \Phi = \frac{R}{Z}$$

From the phasor diagram, we can draw the Impedance triangle as shown in figure.

We have Power, $P = VICos\Phi = [IZ]I \left[\frac{R}{Z} \right] = I^2 R$

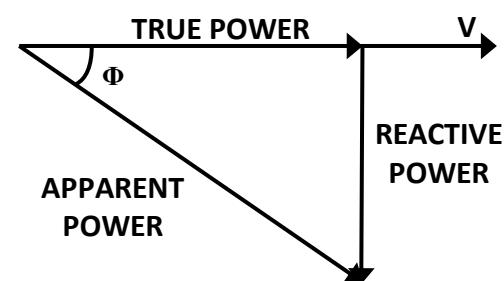
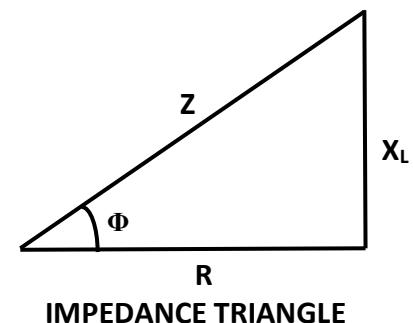
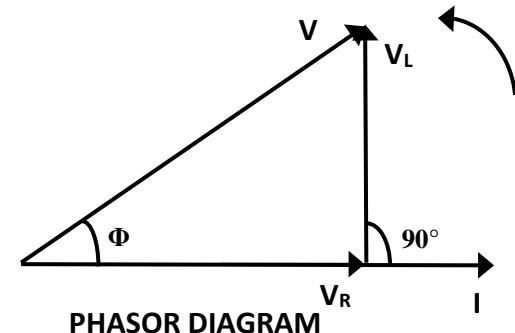
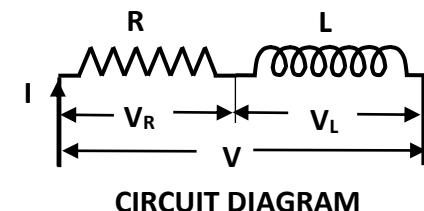
$\therefore \text{Power} = I^2 R$

From the power triangle, we have

True power = $P = VICos\Phi$

Reactive power = $P = VISin\Phi$

Apparent power = $P = VI$



R - C series circuit:

Let us consider a resistor and capacitor in series. If V is the rms value of the applied voltage then I will be the rms value of the current drawn by the circuit.

The voltage across $R = V_R = IR$, where V_R is in phase with I

The voltage across $C = V_C = IX_C$, where V_C lags I by 90°

The applied voltage V is the phasor sum of the two voltage drops V_R and V_C . From the phasor diagram, we have

$$V = \sqrt{V_R^2 + V_C^2} = \sqrt{IR^2 + IX_C^2} = I\sqrt{R^2 + X_C^2}$$

$$I = \frac{V}{\sqrt{R^2 + X_C^2}} \quad \text{or} \quad I = \frac{V}{Z} ; \quad \text{where } Z = \sqrt{R^2 + X_C^2}$$

$$\text{From the phasor diagram, } \tan\Phi = \frac{V_C}{V_R} = \frac{IX_C}{IR} = \frac{X_C}{R}$$

$$\text{or } \Phi = \tan^{-1} \frac{X_C}{R}$$

From the phasor diagram, we can draw the Impedance triangle as shown in figure.

$$\text{From the impedance triangle } \cos\Phi = \frac{R}{Z}$$

$$\text{We have } P = VIC\oslash\Phi \quad \text{or} \quad P = I^2R$$

R - L - C Series circuit:

Let us consider a resistor, inductor and capacitor in series.

If V is the rms value of the applied voltage then I will be the rms value of the current drawn by the circuit.

The voltage across $R = V_R = IR$, where V_R is in phase with I

The voltage across $L = V_L = IX_L$, where V_L leads I by 90°

The voltage across $C = V_C = IX_C$, where V_C lags I by 90°

In the phasor diagram the voltages V_L and V_C are 180° out of phase with each other. V_L is greater in magnitude than V_C so the resultant voltage will be $(V_L - V_C)$.

The applied voltage V will be the phasor sum of the voltages V_R and $(V_L - V_C)$.

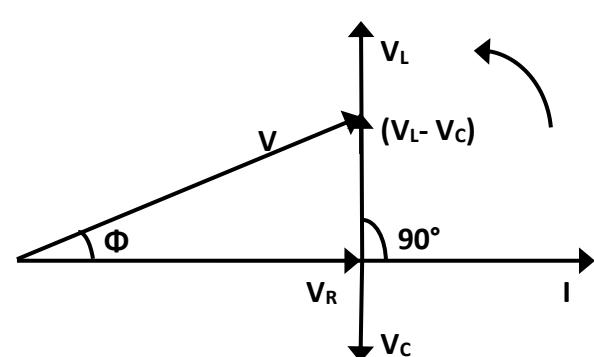
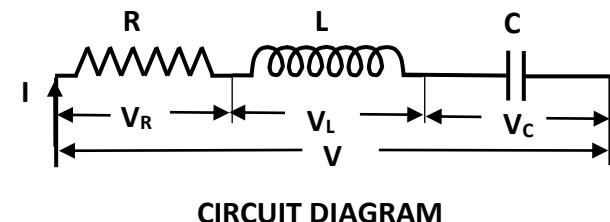
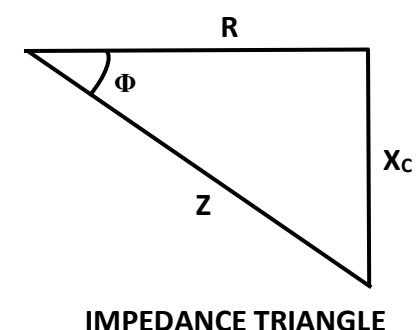
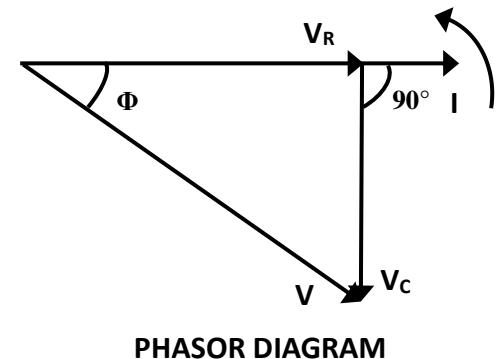
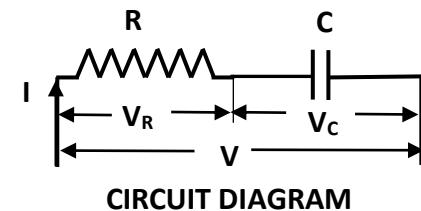
$$\text{We have } V = \sqrt{V_R^2 + (V_L - V_C)^2} = \sqrt{IR^2 + (IX_L - IX_C)^2}$$

$$= I\sqrt{R^2 + (X_L - X_C)^2} \quad \text{or} \quad I = \frac{V}{\sqrt{R^2 + (X_L - X_C)^2}}$$

$$\text{Or } I = \frac{V}{Z} \quad \text{where } Z = \sqrt{R^2 + (X_L - X_C)^2}$$

$$\text{From the phasor diagram, } \tan\Phi = \frac{(V_L - V_C)}{V_R}$$

$$= \frac{(IX_L - IX_C)}{IR} = \frac{(X_L - X_C)}{R} \quad \text{or} \quad \Phi = \tan^{-1} \frac{(X_L - X_C)}{R}$$

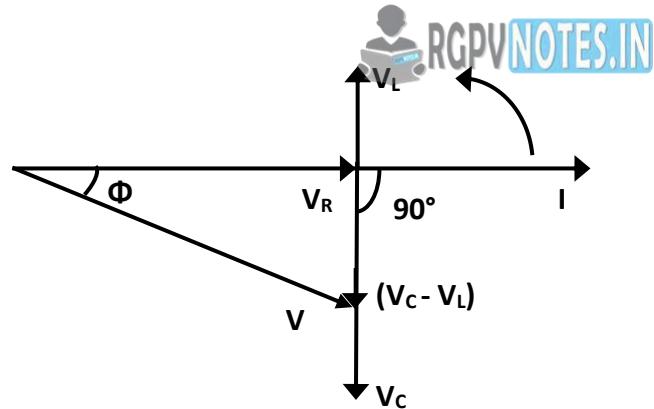
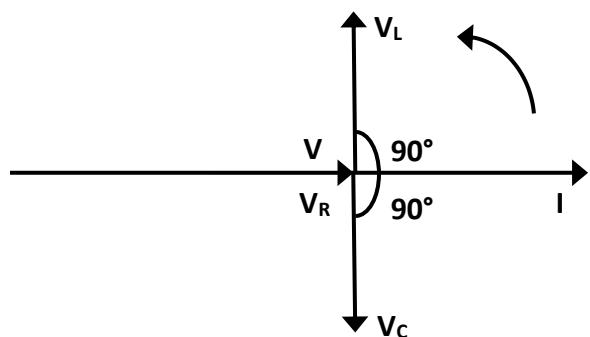


We have $\text{Cos}\Phi = \frac{R}{Z}$ and $P = V I \text{Cos}\Phi$ or $P = I^2 R$

a) For the condition $X_L > X_C$ from the phasor diagram we find that the current lags the applied voltage by an angle Φ , which is greater than zero but less than 90° . So a series RLC circuit with $X_L > X_C$ behaves as a R-L series circuit.

b) For the condition $X_C > X_L$ from the phasor diagram we find that the current leads the applied voltage by an angle Φ which is greater than zero but less than 90° . So a series RLC circuit with $X_C > X_L$ behaves as a R-C series circuit.

c) For the condition $X_L = X_C$ from the phasor diagram we find that the current is in phase with the applied voltage, the phase angle between the current and applied voltage is zero. So the series RLC circuit with $X_L = X_C$ behaves as a pure R circuit. Hence a series R-L-C circuit can behave in three different ways depending upon the values of the Inductive and Capacitive reactances.

PHASOR DIAGRAM FOR $X_L > X_C$ PHASOR DIAGRAM FOR $X_C > X_L$

Three phase AC Circuits - Syllabus

Necessity and advantages of three phase systems, Meaning of Phase sequence, balanced and unbalanced supply and loads. Relationship between line and phase values for balanced star and delta connections. Power in balanced and unbalanced three-phase systems and their measurement. Expression for power factor

THREE PHASE CIRCUITS

Necessity of three phase systems:

Three phase power is usually generated, transmitted and distributed as it has a large number of advantages. Three phase systems are widely used by electrical grids all over the world to transfer power. They are also used to power large motors and other heavy loads. Three phase systems are always adopted because they are very economical.

Advantages of three phase systems:

The advantages of three phase systems over single phase systems are:-

- ❖ Three phase apparatus are smaller in size and lighter in weight than a single phase apparatus with same power output, which makes them cheaper.
- ❖ Three phase systems require only 75% of the weight of conducting material of that required by single phase systems to transmit the same amount of power.
- ❖ Parallel operation of three phase generators is simple when compared to that of single phase generators
- ❖ Output of a three phase machine is 1.5 times the output of a single phase machine of the same size.
- ❖ Three phase motors are self-starting but single phase motors are not self starting
- ❖ Three phase motors have better power factor and efficiency compared to single phase motors.
- ❖ In a single phase circuit, the power delivered is pulsating, whereas, in a three phase system, constant power is delivered when the loads are balanced.
- ❖ In a single phase system, the instantaneous power is not constant and is sinusoidal, which results in vibrations, but in a three phase system, the instantaneous power is always the same.
- ❖ Single phase supply can be managed from a three phase supply, but it is not possible to get a three phase supply from a single phase supply.
- ❖ Three phase supply can be rectified into dc supply with lesser ripple factor.

Phase sequence: It is the order in which the voltages in the three coils reach their positive maximum values one after the other.

Balanced systems:

In a three phase circuit if all the impedances are equal then the circuit is called a balanced system. To a balanced circuit if balanced three phase voltage is applied the currents flowing shall also be balanced. The three phase voltages are equal in magnitude but 120° out of phase from each other and so are the three currents. The sum of all the phase currents or the phase voltages will be equal to zero.

Unbalanced systems:

The three phase systems can be unbalanced. The unbalance may be due to the unbalance in the supply or unbalance in the load. Sometimes both the supply and the load may be unbalanced.

In the analysis of unbalanced systems each phase has to be treated separately and the resultant can be obtained by their phasor representation. Let us consider a balanced supply system where the three phase voltages are : $V \angle 0^\circ$, $V \angle 120^\circ$, $V \angle 240^\circ$. Let the unbalanced impedances be : $Z_1 \angle \theta_1^\circ$, $Z_2 \angle \theta_2^\circ$, $Z_3 \angle \theta_3^\circ$ respectively in the three phases.

The current in each phase will be:

$$I_1 = \frac{V}{Z_1} \angle -\theta_1^\circ; \quad I_2 = \frac{V}{Z_2} \angle 120 - \theta_2^\circ; \quad I_3 = \frac{V}{Z_3} \angle 240 - \theta_3^\circ$$

Star connection: If similar ends of the three phase windings are joined together at a common point N, a star connection is obtained.

Delta connection: If dissimilar ends of the three phase windings are joined to form a closed loop, a delta connection is obtained.

Relations in star connection:

From the figure - 1, it is observed that, the current flowing in the phase winding of each phase = The current flowing in that respective line. Hence, Phase current = Line current or $I_{PH} = I_L$

Consider the lines R and Y, line voltage V_{RY} is the phasor difference of E_{RN} and E_{YN} . To subtract E_{YN} from E_{RN} , the phasor E_{YN} is reversed and the phasor sum with E_{RN} is obtained. The two phasors E_{RN} and $-E_{YN}$ are equal in magnitude and equal to E_{PH} and are 60° apart as observed in the phasor diagram shown in figure - 2.

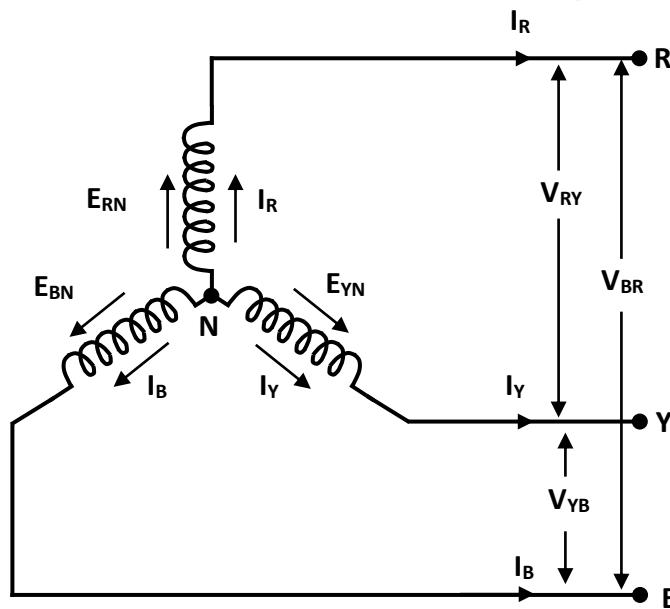


Fig. - 1

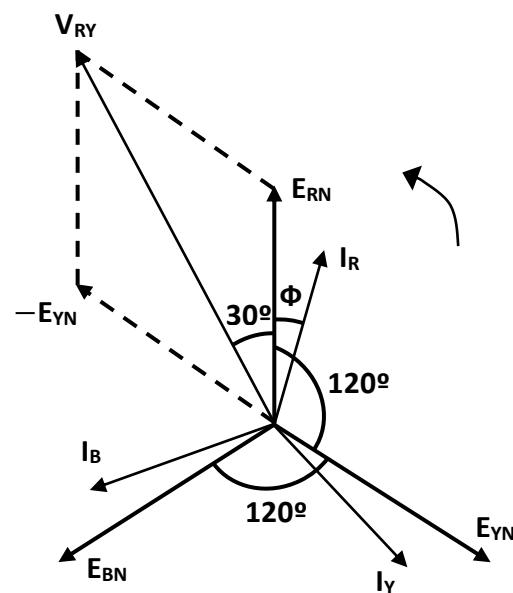


Fig. - 2

From the phasor diagram shown in figure - 2, we have

$$V_{RY} = 2E_{PH} \cos\left(\frac{60}{2}\right) = 2E_{PH} \cos 30^\circ = \sqrt{3}E_{PH}$$

Similarly $V_{YB} = E_{YN} - E_{BN} = \sqrt{3}E_{PH}$ and $V_{BR} = E_{BN} - E_{RN} = \sqrt{3}E_{PH}$

$$\therefore V_L = \sqrt{3}V_{PH} \text{ and } I_L = I_{PH}$$

From figure - 2, it is observed that: The line voltages are 120° apart

The line voltages are 30° ahead of their respective phase voltages

The angle between the line currents and corresponding line voltages is $(30 + \Phi)$

Power per phase = $V_{PH}I_{PH}\cos\phi$; Total power = $3V_{PH}I_{PH}\cos\phi$

With line values, Total power = $\sqrt{3}V_LI_L\cos\phi$ where ϕ is the phase angle between V_{PH} and I_{PH}

Relations in delta connection:

From the figure - 3, it is observed that one phase winding is included between any pair of lines.

Hence, the Line voltage = Phase voltage ie. $V_L = V_{PH}$

The current in any line is equal to the phasor difference of the currents in the two phases attached to that line. Hence, the current in line R is the phasor difference of I_R and I_B . To subtract I_B from I_R , the phasor I_B is reversed and its phasor sum with I_R is obtained. The two phasors I_R and $-I_B$ are equal in magnitude and equal to I_{PH} and are 60° apart as observed in the phasor diagram shown in figure - 4.

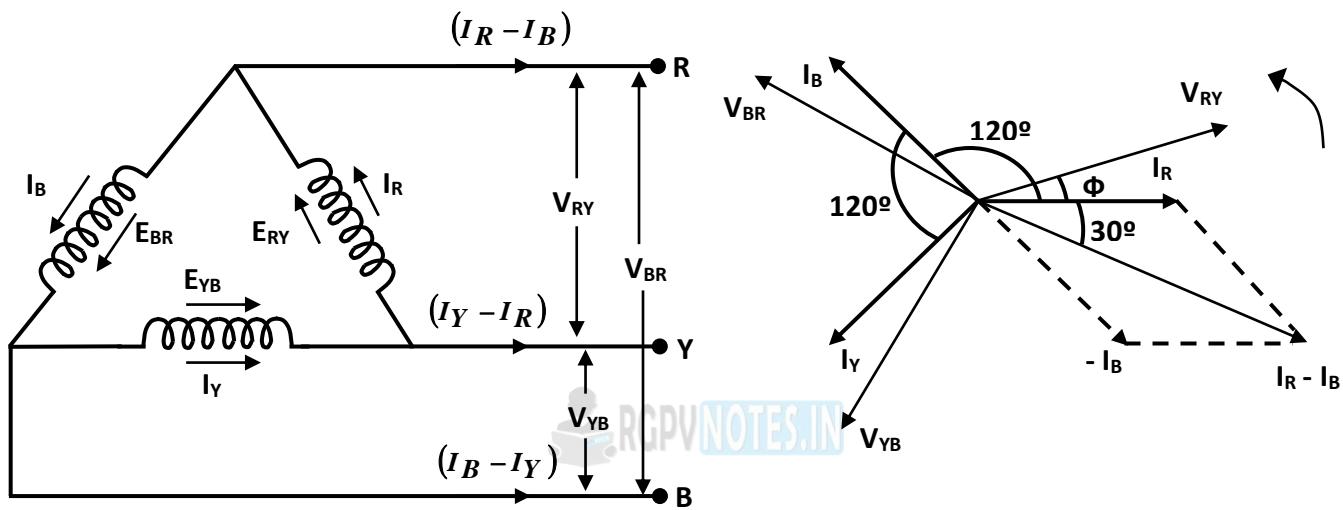


Fig. - 3

Fig. - 4

From the phasor diagram shown in figure - 4, we have

$$(I_R - I_B) = 2I_{PH}\cos\left(\frac{60}{2}\right) = 2I_{PH}\cos30^\circ = \sqrt{3}I_{PH}$$

Similarly $(I_Y - I_R) = \sqrt{3}I_{PH}$ and $(I_B - I_Y) = \sqrt{3}I_{PH}$

$$\therefore I_L = \sqrt{3}I_{PH} \text{ and } V_L = V_{PH}$$

From figure - 4, it is observed that: The line currents are 120° apart

The line currents are 30° behind their respective phase currents

The angle between the line currents and corresponding line voltages is $(30 + \Phi)$

Power per phase = $V_{PH}I_{PH}\cos\phi$; Total power = $3V_{PH}I_{PH}\cos\phi$

With line values, Total power = $\sqrt{3}V_LI_L\cos\phi$ where ϕ is the phase angle between V_{PH} and I_{PH}

Measurement of three phase power using two wattmeter for Star connected load:

Let us consider the loads to be connected in star as shown in figure - 5. The current coils of the two wattmeters are connected in line R and line B. The potential coils of the wattmeters are connected across lines R and Y as well as lines B and Y.

Let the instantaneous values of potential difference across the loads be v_r , v_y , v_b and the corresponding values of instantaneous line currents be i_r , i_y , i_b

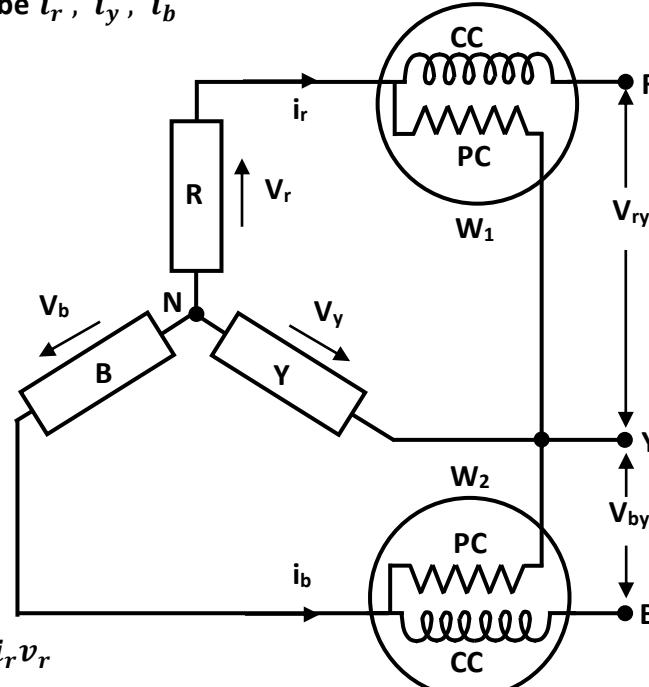


Fig. - 5

The instantaneous power in load $R = i_r v_r$

The instantaneous power in load $Y = i_y v_y$

The instantaneous power in load $B = i_b v_b$

Total instantaneous power = $i_r v_r + i_y v_y + i_b v_b$

From the figure - 5, it is observed that the instantaneous current through current coil of $W_1 = i_r$ and the instantaneous potential difference across its potential coil = $(v_r - v_y)$

Instantaneous power measured by $W_1 = i_r(v_r - v_y)$

Similarly the instantaneous current thro current coil of $W_2 = i_b$ and the instantaneous potential difference across its potential coil = $(v_b - v_y)$

Instantaneous power measured by $W_2 = i_b(v_b - v_y)$

$$W_1 + W_2 = i_r(v_r - v_y) + i_b(v_b - v_y)$$

$$\text{or } W_1 + W_2 = i_r v_r + i_b v_b - v_y(i_r + i_b) \quad \text{--- (1)}$$

Applying KCL to the junction N shown in figure - 5, we get

$$i_r + i_y + i_b = 0 \quad \text{or} \quad (i_r + i_b) = -i_y \quad \text{--- (2)}$$

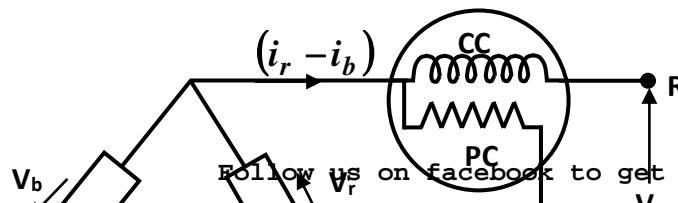
Introducing equation (2) in equation (1) we get,

$W_1 + W_2 = i_r v_r + i_y v_y + i_b v_b$ which is equal to total instantaneous power consumed by a three phase star connected load.

Measurement of three phase power using two wattmeter for Delta connected load:

Let us consider the loads to be connected in delta as shown in figure - 6.

Let the instantaneous values of potential difference across the loads be v_r , v_y , v_b and the corresponding values of instantaneous phase currents be i_r , i_y , i_b



Follow us on facebook to get real-time updates from RGPNOTES

The instantaneous power in load R = $i_r v_r$

The instantaneous power in load Y = $i_y v_y$

The instantaneous power in load B = $i_b v_b$

Total instantaneous power = $i_r v_r + i_y v_y + i_b v_b$

From the figure - 6, it is observed that the instantaneous current through current coil of $W_1 = (i_r - i_b)$ and the instantaneous potential difference across its potential coil = v_r

Instantaneous power measured by $W_1 = (i_r - i_b)v_r$

Similarly the instantaneous current through current coil of $W_2 = (i_b - i_y)$ and the instantaneous potential difference across its potential coil = $(-v_y)$

Instantaneous power measured by $W_2 = (i_b - i_y)(-v_y)$

$$W_1 + W_2 = (i_r - i_b)v_r + (i_b - i_y)(-v_y)$$

$$\text{or } W_1 + W_2 = i_r v_r + i_y v_y - i_b(v_r + v_y) \quad \dots (1)$$

Applying KVL to closed loop ABC in figure - 6, we get

$$v_r + v_y + v_b = 0 \quad \text{or} \quad (v_r + v_y) = -v_b \quad \dots (2)$$

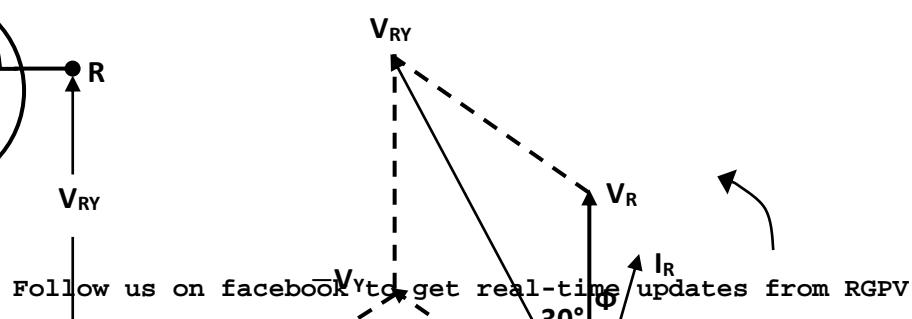
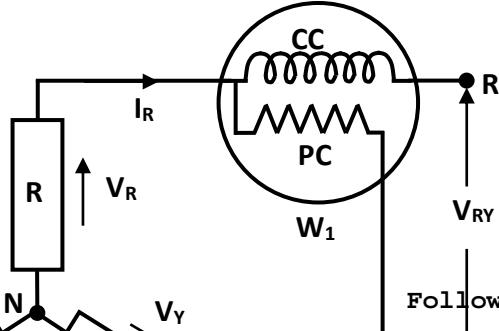
Introducing equation (2) in equation (1), we get

$$W_1 + W_2 = i_r v_r + i_y v_y - i_b(-v_b)$$

$\therefore W_1 + W_2 = i_r v_r + i_y v_y + i_b v_b$ which is equal to total instantaneous power consumed by a three phase delta connected load.

Expression for power factor in terms of wattmeter readings:

Let us consider a three phase balanced star connected load with a lagging phase angle Φ . Let V_R , V_Y and V_B be the rms values of phase voltages across the star connected load and I_R , I_Y and I_B be the phase currents.



Since the load has a lagging power factor, the phase currents lag their respective phase voltages by an angle Φ as shown in the phasor diagram. The power is measured using two wattmeters. The current through current coil of $W_1 = I_R$ and the potential difference across its potential coil = V_{RY} .

From the phasor diagram the phase angle between V_{RY} and $I_R = (30 + \Phi)$

\therefore Reading of $W_1 = I_R V_{RY} \cos (30 + \Phi)$

The current through current coil of $W_2 = I_B$ and the potential difference across its potential coil = V_{BY} .

From the phasor diagram the phase angle between V_{BY} and $I_B = (30 - \Phi)$.

\therefore Reading of $W_2 = I_B V_{BY} \cos (30 - \Phi)$

Since the load is balanced $I_R = I_Y = I_B = I_L$ and $V_{RY} = V_{BY} = V_{BR} = V_L$

$\therefore W_1 = V_L I_L \cos (30 + \Phi)$ and $W_2 = V_L I_L \cos (30 - \Phi)$

$$W_2 + W_1 = V_L I_L [\cos 30 \cos \Phi + \cancel{\sin 30 \sin \Phi} + \cos 30 \cos \Phi - \cancel{\sin 30 \sin \Phi}]$$

$$W_2 + W_1 = V_L I_L (2 \cos 30 \cos \Phi) = V_L I_L (2 \frac{\sqrt{3}}{2} \cos \Phi) = \sqrt{3} V_L I_L \cos \Phi$$

$\therefore W_2 + W_1 = \text{Total Three phase power. } W_2 + W_1 = \sqrt{3} V_L I_L \cos \Phi \text{ ---- (1)}$

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

$$= V_L I_L [\cancel{\cos 30 \cos \Phi} + \sin 30 \sin \Phi - \cancel{\cos 30 \cos \Phi} + \sin 30 \sin \Phi]$$

$$= V_L I_L [2 \sin 30 \sin \Phi] = V_L I_L [2 \frac{1}{2} \sin \Phi] = V_L I_L \sin \Phi \text{ ---- (2)}$$

Dividing equation (2) by equation (1) we get

$$\frac{(W_2 - W_1)}{(W_2 + W_1)} = \frac{(V_L I_L \sin \phi)}{(\sqrt{3} V_L I_L \cos \phi)} = \frac{\tan \phi}{\sqrt{3}} \text{ or } \therefore \tan \phi = \sqrt{3} \frac{(W_2 - W_1)}{(W_2 + W_1)}$$

$$\text{or } \phi = \tan^{-1} \left(\sqrt{3} \frac{(W_2 - W_1)}{(W_2 + W_1)} \right) \quad \therefore \cos \phi = \cos \left(\tan^{-1} \sqrt{3} \frac{(W_2 - W_1)}{(W_2 + W_1)} \right)$$

Variation in wattmeter readings for lagging power factors:

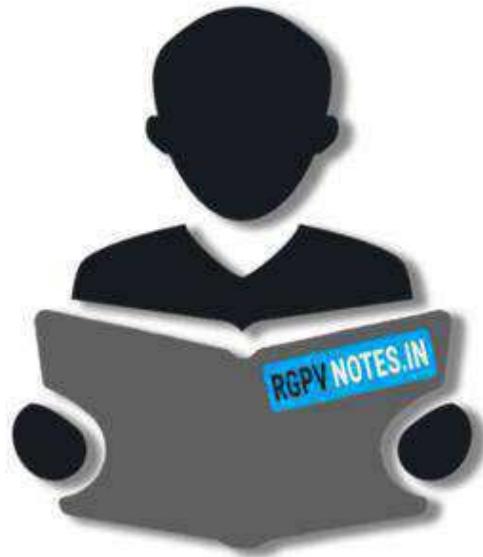
The wattmeter readings for lagging power factors will be -

$$W_1 = V_L I_L \cos (30 + \Phi) \quad \text{and} \quad W_2 = V_L I_L \cos (30 - \Phi)$$

The phase angle Φ can be assumed to possess different values and the variation of the wattmeter readings can be observed.

Phase angle Φ	0°	60°	90°
Wattmeter reading W_1	+ ve	0	- ve
Wattmeter reading W_2	+ ve	+ ve	+ ve
	$W_1 = W_2$ for resistive loads		$W_1 = W_2$





RGPVNOTES.IN

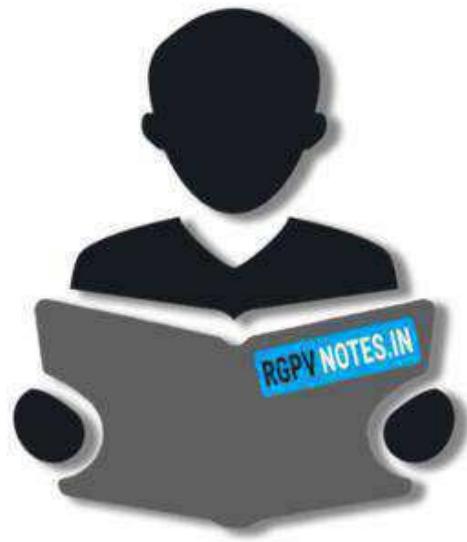
We hope you find these notes useful.

You can get previous year question papers at
<https://qp.rgpvnotes.in> .

If you have any queries or you want to submit your
study notes please write us at
rgpvnotes.in@gmail.com



LIKE & FOLLOW US ON FACEBOOK
facebook.com/rgpvnotes.in



RGPVNOTES.IN

Program : **B.Tech**

Subject Name: **Engineering Chemistry**

Subject Code: **BT-101**

Semester: **1st**



LIKE & FOLLOW US ON FACEBOOK
facebook.com/rgpvnotes.in

UNIT II: BOILER PROBLEMS & SOFTENING METHODS

BOILER WATER (WATER FOR STEAM GENERATION)

A boiler is a closed vessel in which water under pressure is transferred into steam by the application of heat. In the boiler furnace, the chemical energy in the fuel is converted into heat, and it is the function of the boiler to transfer this heat to the contained water in the most efficient manner. The boiler should also be designed to generate high quality steam for plant use. A boiler must be designed to absorb the maximum amount of heat released in the process of combustion. This heat is transferred to the boiler water through radiation, conduction and convection.

Steam utilization. Steam is generated for the following plant uses:

- (i) Turbine drive for electric generating equipment, blowers and pumps,
- (ii) Heating for direct contact for equipment and comfort,
- (iii) Process for direct contact with products; direct contact sterilization and noncontact for processing temperatures.

Water is mainly used in boilers for the generation of steam (for industries and power houses). For such water all the impurities are not necessarily eliminated, and only those impurities which lead to operational troubles in boilers are eliminated or kept within the tolerable limits.

Boiler-feed water should correspond with the following composition:

- (i) Its hardness should be below 0.2 ppm.
- (ii) Its caustic alkalinity (due to OH⁻) should be between 0.15 and 0.45 ppm.
- (iii) Its soda alkalinity (due to Na₂CO₃) should be 0.45 – 1 ppm.

Excess of impurities, if present, in boiler feed water generally cause the following problems:

Scale and sludge formation, corrosion, priming and foaming, caustic embrittlement.

BOILER PROBLEM:

1. SLUDGE & SCALE FORMATION
2. PRIMING & FOAMING
3. CARRY OVER
4. BOILER CORROSION
5. CAUSTIC EMBRITTLEMENT

1. Sludge and Scale Formation in Boilers

In a boiler, water is continuously evaporated to form steam. This increases the concentration of dissolved salts. Finally a stage is reached when the ionic product of these salts exceeds their solubility product and hence they are thrown out as precipitates.

If the precipitates formed are soft loose and slimy, these are known as *sludges*, while if the precipitate is hard and adhering on the inner walls, it is called as *scale*.

SULDGE: Sludge is a soft, loose and slimy precipitate formed within the boiler. Sludges are formed by substances which have greater solubility in hot water than in cold water, e.g. MgCO₃, MgCl₂, CaCl₂, MgSO₄ etc. They are formed at comparatively colder portions of the boiler get collected at places where the flow rate is slow; they can be easily removed (scrapped off) with a wire brush. If sludges are formed along with scales, then former gets entrapped in the latter and both get deposited as scales.

Disadvantages of sludge formation:-

- (i) Sludges are poor conductors of heat, so they tend to waste a portion of heat generated and thus decrease the efficiency of boiler.
- (ii) Excessive sludge formation disturbs the working of the boiler. It settles in the regions of poor water circulation such as pipe connection, plug opening, gauge-glass connection, thereby causing even choking of the pipes.

Prevention of sludge formation:-

- (i) By using softened water
- (ii) By frequently '**blow-down operation**', (i.e. partial removal of concentrated water through a tap at the bottom of boiler, when extent of hardness in the boiler becomes alarmingly high).

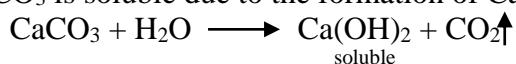
SCALES: Scales are hard deposits firmly sticking to the inner surfaces of the boiler. They are difficult to remove, even with the help of hammer and chisel, and are the main source of boiler troubles.

(i) Decomposition of calcium bicarbonate:-



However, scale composed chiefly of calcium carbonate is soft and is the *main cause of scale formation in low-pressure boilers*.

But in high-pressure boilers, CaCO_3 is soluble due to the formation of $\text{Ca}(\text{OH})_2$



(iii) Deposition of calcium sulphate:-

The solubility of CaSO_4 in water decreases with increase in temperature. CaSO_4 is soluble in cold water, but almost completely insoluble in super-heated water. It may be due to increase ionization at high temperature so $k_{sp} < k_{\text{ionic prod.}}$ and less availability of water molecules for solvation at high temperature.

Consequently, CaSO_4 gets precipitated as hard scale on the hotter parts, of the boiler. *This type of scale causes troubles mainly in high pressure boilers.* Calcium sulphate scale is quite adherent and difficult to remove, even with the help of hammer and chisel.

(iii) Hydrolysis of magnesium salts

Dissolved magnesium salts get hydrolyzed (at prevailing high temperature inside the boiler) forming magnesium hydroxide precipitate, which forms a soft type of scale, e.g.



(iv) Presence of silica:-

Even if a small quantity of SiO_2 is present, it may deposit as calcium silicate (CaSiO_3) and / or magnesium silicate (MgSiO_3). These deposits adhere very firmly on the inner side of the boiler surface and are very difficult to remove. One important source of silica in water is the sand filter.

Disadvantages of Scale formation:

(i) *Wastage of fuel.* Scales have a poor thermal conductivity so the rate of heat transfer from boiler to inside water is greatly reduced. In order to provide a steady supply of heat to water, excessive or over-heating is done and these causes increase in fuel consumption.

The wastage of fuel being dependent on the thickness and the nature of scale:

<i>Thickness of scale (mm)</i>	0.325	0.625	1.25	2.5	12
<i>Wastage of fuel</i>	10%	15%	50%	80%	150%

(ii) *Lowering of boiler safety.* Due to scale formation, over-heating of boiler is done in order to maintain a steady supply of steam. It makes the boiler material softer and weaker. This causes distortion of boiler tube and also makes the boiler unsafe to bear the pressure of the steam, especially in high-pressure boilers.

(iii) *Decrease in efficiency.* Deposition of scales in the valves and condensers of the boiler, choke them partially. This results in decrease in efficiency of the boiler.

(iv) *Danger of explosion.* When thick scales crack due to uneven expansion, the water comes suddenly in contact with over-heated portion and large amount of steam is formed instantaneously. This results in development of sudden or sudden high-pressure which may cause explosion of the boiler.

Removal of Scales:

Scales are removed by mechanical methods (i – iii) and / or by chemical methods (iv)

(i) If the scales are loosely adhering, it can be removed with the help of scraper or piece of wood or wire brush,

(ii) If the scales are brittle, it can be removed by giving thermal shocks (i.e., heating the boiler and then suddenly cooling with cold water).

(iii) If the scales are loosely adhering, they can also be removed by frequent blow-down operation.

Blow-down operation is partial removal of hard water through a ‘tap’ at the bottom of the boiler, when extent of hardness in the boiler becomes alarmingly high. ‘Make-up’ water is addition of fresh softened water to boiler after blow down operation.

(iv) If the scales are adherent and hard, they can be removed by dissolving them by adding chemicals e.g., CaCO_3 scales can be dissolved by using 5-10% HCl. Calcium sulphate scales can be removed by adding EDTA, since the Ca – EDTA complex is highly soluble in water.

The essential differences between sludges and scales are summarized as follows:

S.No.	Sludges	Scales
1.	Sludges are soft, loose and slimy precipitate.	Scales are hard deposits.
2.	They are non-adherent deposits and can be easily removed.	They stick very firmly to the inner surface of boiler and are very difficult to remove.
3.	Formed by substances like CaCl_2 , MgCl_2 , MgSO_4 , MgCO_3 etc.	Formed by substance like CaSO_4 , Mg(OH)_2 etc.
4.	Formed at comparatively colder portions of the boiler.	Formed generally at heated positions of the boiler.
5.	They decrease the efficiency of boiler but are less dangerous.	Decrease the efficiency of boiler and chances of explosions are also there.
6.	Can be removed by blow-down operation.	Cannot be removed by blow-down operation.

Priming and Foaming

When steam is produced rapidly in the boilers, some droplets of the liquid water are carried along-with the steam. This process of ‘wet-steam’ formation is called *priming*.

Priming refers to the propulsion of water into the steam drum by extremely rapid, almost explosive boiling of the water at the heating surfaces.

The moisture contamination in the steam is expressed in percentage by weight of steam.

For example, if steam contains 0.2% moisture, its steam quality will be reported as $100 - 0.2 = 99.8\%$.

Priming is caused by:

- (i) The presence of considerable quantities of dissolved solids (mainly due to suspended impurities and due to dissolved impurities in water).
- (ii) Steam velocities high enough to carry droplets of water into the steam pipe;
- (iii) Sudden boiling;
- (iv) Faulty design of boiler.

Priming can be avoided by:

- (i) Controlling rapid change in steaming velocities,
- (ii) The proper design of boilers (maintaining low water levels in boilers)
- (iii) Ensuring efficient softening and
- (iv) Filtration of the boiler-water carried over to the boiler.
- (v) By blowing off sludge or scales from time to time.

Foaming is the formation of small but persistent foam or bubbles at the water surface in boilers, which do not break easily. *Foaming is caused by* the presence of oil and alkalis in boiler-feed water. Actually oils and alkalis react to form scaps, which greatly lowers the surface tension of water, and thus increase the foaming tendency of the liquid.

With respect to foaming, water can be following grades:

- (i) *Foaming water.* It is that water which produces foam even in two days, if blowing off operation is not done.
- (ii) *Semi-Foaming water.* It is that water which does not produce any foam in locomotive boilers for two days.
- (iii) *Non-Foaming water.* It is that water which does not produce any foam in locomotive boilers for one week.

Foaming can be avoided by: (i) the addition of anti-foaming agents, which act by count enacting the reduction in surface tension. For example addition of castor oil (which spreads on the surface of water and therefore) neutralizes the surface tension reduction. (ii) The removal of foaming agent (oil) from boiler water.

Traces oils are generally introduced in boiler feed water through the lubricating materials used for pumps etc. Oils can be removed by the addition of aluminum compounds, like *sodium aluminate* and *aluminium; sulphate* which is hydrolyzed to form aluminium hydroxide flocks which entrap oil drops. The flocks of $\text{Al}(\text{OH})_3$ containing oil droplets are removed by filtration through anthracite filter bed.

Carry Over: The phenomenon of carrying of water along with impurities by steam is called “carry over”. This is mainly due to priming and foaming.

Priming and foaming usually occur together.

They are objectionable, usually occur together.

- (i) Dissolved salts or suspended solids in boiler water are carried by the wet steam to super heater and turbine blades, where they get deposited as water evaporates. This deposit decreases the efficiency of boiler.
- (ii) Dissolved salts may enter the parts of other machinery, thereby decreasing their life;
- (iii) The maintenance of the boiler pressure becomes difficult because of improper judgment of actual height of water column.

Boiler Corrosion

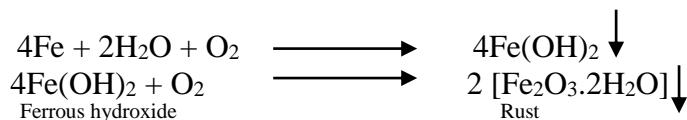
Boiler corrosion is “decay” or “disintegration” of boiler body material either due to chemical or electrochemical reaction with its environment.

The disadvantages of corrosion are:

- (i) Shortening of boiler life,
- (ii) Leaks of the joints and rivets;
- (iii) Increased cost of repairs and maintenance

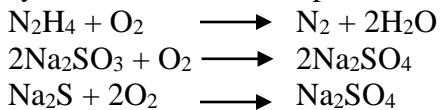
Corrosion in boilers is due to the following reasons:

(1) **Dissolved oxygen.** This is the most usual corrosion causing factor. In Boilers, oxygen is introduced through the raw water supply. Water usually contains about 8 ppm of dissolved oxygen at room temp. As the water is heated, the dissolved oxygen is set free and the boiler starts corroding. Dissolved oxygen reacts with the iron of boiler in presence of water and under prevailing high temperature to form ferric oxide (rust).



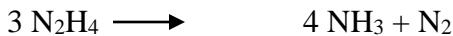
Removal of dissolved oxygen:

- (i) By adding hydrazine or sodium sulphate or sodium sulphide. Thus:



Hydrazine is an ideal chemical for the removal of dissolved oxygen. It reacts with oxygen, forming nitrogen and water. Nitrogen is harmless. Consequently hydrazine removes oxygen without increasing the conc. of dissolved solids/salts.

- (a) Pure hydrazine is not used in water treatment because it is an explosive inflammable liquid so 40% aqueous solution of hydrazine is used which is quite safe.
- (b) Excess hydrazine must not be used because excess of it decomposes to give NH_3 , which causes corrosion of some alloys like brass etc. used in condenser tubes.



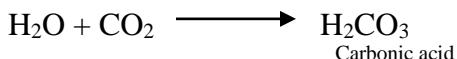
On the other hand, if sodium sulphite or sodium sulphide is used, the sodium sulphate is formed. Under high pressure it decomposes giving SO_2 . The SO_2 enters the steam pipes and appears as corrosive sulphurous acid (H_2SO_3) in steam condensate. So as a rule a very low concentration of 5-10 ppm of Na_2SO_3 in the boiler is maintained, rather adding it intermittently.

- (ii) By mechanical de-aeration. This process consists of spraying water over preheated perforated plates stacked in a degasifier. Removal of dissolved O_2 is ensured by applying high temperature and vacuum.

(2) **Carbon dioxide.** There are two sources of CO_2 in boiler water, viz. dissolved CO_2 in raw water and CO_2 formed by decomposition of bicarbonates in H_2O according to the equation:



Carbon dioxide in presence of water forms carbonic acid which has a corrosive effect on the boiler material like any other acid.



CO_2 can be removed by:

- (i) Mechanical de-aeration along with O_2 .

(ii) Filtering water through lime-stone



But this method increases hardness

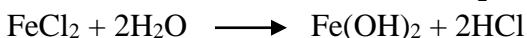
(iii) Addition of appropriate quantity of ammonium hydroxide



(3) **Mineral acids.** Magnesium chloride, if present in boiler feed water, can undergo hydrolysis producing HCl



The liberated acid reacts with iron material of the boiler to form ferrous hydroxide which in turn is converted to rust in the following way:



Thus, a small amount of HCl may cause extensive corrosion since HCl is produced in a chain-like manner. Consequently presence of even a small amount of MgCl₂ causes corrosion of iron to a large extent.

As the boiler water is generally alkaline and hence the acid is usually neutralized. In case the amount of acid is more, calculated quantity of alkali is added from outside to neutralize the acid for preventing this corrosion.

Caustic Embrittlement

Caustic embrittlement is the phenomenon during which the boiler material becomes brittle due to the accumulation of caustic substances. This type of boiler corrosion is caused by the use of highly alkaline water in the high pressure boiler.

During softening by lime-soda process, it is likely that some residual Na₂CO₃ is still present in the softened water. In high pressure boilers Na₂CO₃ decomposes to give sodium hydroxide and CO₂, and sodium hydroxide thus produced makes the boiler water “caustic”.



This caustic water flows into the minute hair-cracks, present in the inner side of boiler, by capillary action. On evaporation of water the dissolved caustic soda concentration increases progressively which attacks the surrounding area, thereby dissolving iron of boiler as Sodium ferroate (Na₂FeO₂).

From its place of formation, sodium ferroate decomposes a short distance away as per the following equation.



Further dissolution of iron takes place because of

(i) The precipitation of Fe₃O₄, and

(ii) The regeneration of NaOH.

This causes embrittlement of boiler walls more particularly stressed parts (like bends, joints, rivets, etc.), causing even failure of the boiler.

Mechanically embrittlement arises due to the setting up of a *concentration cell*.

With the iron surrounded by *dil. NaOH* acting as the *Cathode*, while the iron surrounded by *conc. NaOH* acting as the *anode*.

The iron in the anodic part gets dissolved or corroded.

Caustic embrittlement can be prevented:

(i) by using sodium phosphate as softening reagent, instead of sodium carbonate in external treatment of boiler water.

(ii) by adding tannin or lignin to boiler water which blocks the hair-cracks in the boiler walls thereby preventing infiltration of caustic soda solution into these areas.

(iii) by adding sodium sulphate to boiler water:

Na₂SO₄ also blocks hair-cracks, thereby preventing infiltration of caustic soda solution in these. It has been observed that caustic cracking can be prevented if Na₂SO₄ is added to boiler water so that the ratio:

$\frac{[\text{Na}_2\text{SO}_4 \text{ conc.}]}{[\text{NaOH conc.}]}$ is kept as 1:1, 2:1 and 3:1 in boilers working respectively at pressures up to 10, 20 and above 20 atmospheres.

Softening Methods

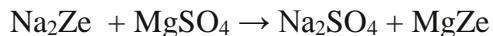
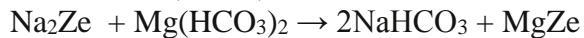
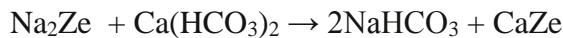
Zeolite process

Zeolite is micro-porous mineral which is used as catalyst in many industrial purposes such as water purification and air purification. The zeolites are hydrated alumino silicates and general composition Al_xSi_yO_{2(x+y)} (without water molecules). Zeolites are two types natural and synthetic or artificial. The natural zeolite that is used for water softening is gluconites or greensand. Permutit is the synthetic zeolite that is most used in water softening and its chemical formula is Na₂O, Al₂O₃, nSiO₂, xH₂O. These are used as ion exchanger and odor removal in water softener. Permutit are more porous, glassy, and have higher softening capacity than greensand.

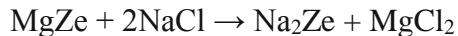
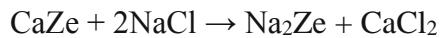
Zeolites are characteristically soft to moderately hard, light in density, insoluble in water but can act as base exchangers in contact with water containing cations. Hence these can remove Ca²⁺ and Mg²⁺ ions from water when hard water is passes through.

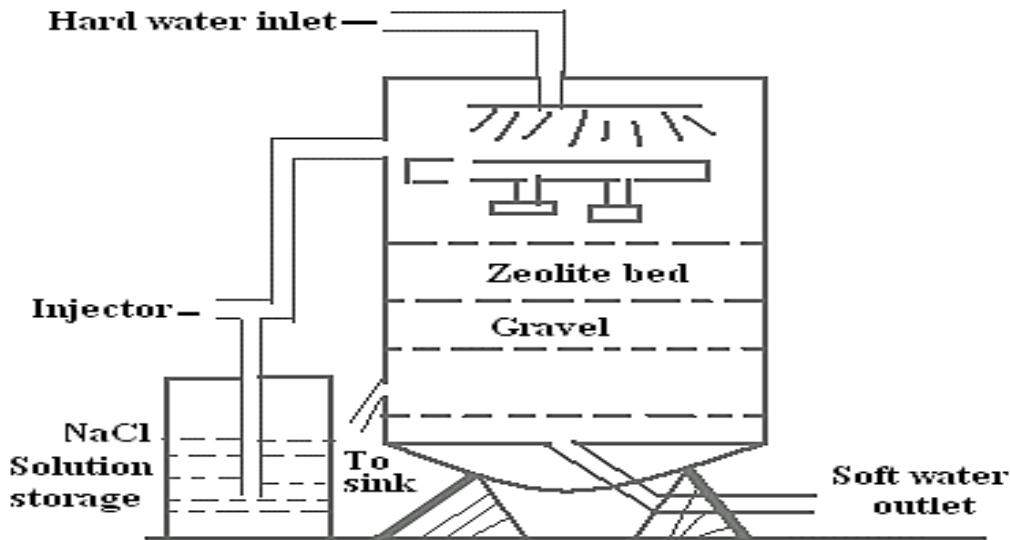
Zeolite process for water softening

Zeolite process for water softening has become a commercial success for the reason that zeolite can be easily regenerated. When Ca²⁺ and Mg²⁺ ions containing hard water is passes through a bed of sodium zeolite, the sodium ions are replace by the calcium and magnesium ions.



When all sodium ions are replaced by calcium and magnesium ions, the zeolite becomes inactive. Then the zeolite needs to be regenerated. Brine solutions are passing through the bed of inactivated zeolite. The following reactions are taken place and form Na₂Ze





Merits of Zeolite Process:

It removes the hardness almost completely (about 10 ppm hardness only).

The process automatically adjust itself for variation in hardness of incoming water.

This process does not involve any type of precipitation, thus, no problem of sludge formation occurs.

Demerits of Zeolite Process:

The outgoing water (treated water) contains more sodium salts.

This method only replaces Ca^{+2} and Mg^{+2} ions by Na^+ ions.

High turbidity water cannot be softened efficiently by zeolite process.

Ion exchange process Or Demineralization process Or Deionization process

Ion exchanger resin are insoluble , cross linked, long chain higher molecular weight organic polymers which are permeable due to their micro porous structure, and the functional groups attached to the chains are involved in the ion-exchanging properties.

Types of ion exchanger resins

Cation exchanger resins (RH^+) : These are usually styrene divinyl benzene copolymers which on carboxylation or sulphonation become capable of exchanging their H^+ ions with the cations of the solution. These have acidic functional groups like SO_3H^+ , $-\text{COOH}$ or OH^-

Anion exchanger resins (ROH^-) : These are usually styrene divinyl benzene or amine formaldehyde copolymers which on treatment with dilute NaOH solution become capable to exchange their OH^- anions with anions in water. Which contains basic functional groups like quaternary ammonium or quaternary sulphonium or quaternary phosphonium groups.

Ion exchange process

It is a two stage process, the water first passes through the column containing hydrogen exchanger (Cation exchanger) and then through second column containing hydroxyl exchanger

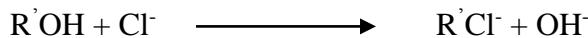
(Anion exchanger). H^+ and OH^- ions released from cation exchanger and anion exchanger columns get combined to produce water molecule.

Cation exchange resins :



(RH^+ = Cation exchange resin)

Anion exchange resins :

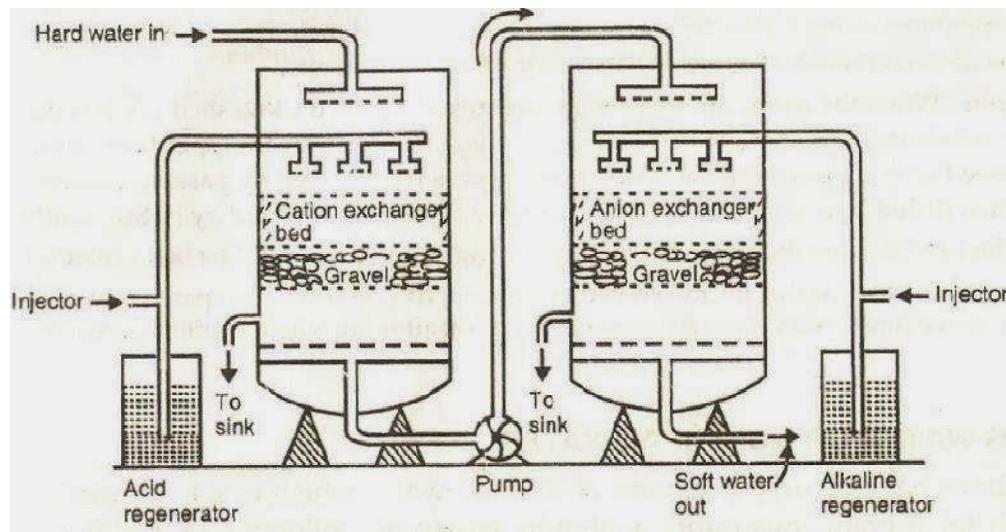


($\text{R}'\text{OH}^-$ = Anion exchange resin)

H^+ + OH^- ions , combine and produce water



Ion exchange process Diagram



Regeneration of exhausted resins

After some time the cation and anion exchanging resins lose the capacity to remove H^+ ions and OH^- ions respectively, they are then said to be exhausted.

When the resins are exhausted, the supply of water is stopped. The exhausted cation exchanger is regenerated by passing dilute HCl or H_2SO_4 Solution and exhausted anion exchanger resin is regenerated by passing dilute NaOH solution.

Regeneration Chemical reaction



The columns are washed with deionised water and washing (which contain Ca^{2+} , Mg^{2+} , SO_4^{2-} ions) are passed to sink or drain. The regenerated resins are again used.

Advantages:

Highly acidic or alkaline water also can be softened by this process.

It produces water of low hardness (up to 2 ppm).

If the output water is passed through de-gassifier, then the gaseous impurities like O_2 , CO_2 also get expelled, to get water of distilled water standard.

Disadvantages:

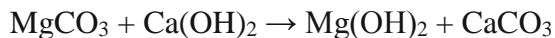
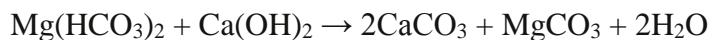
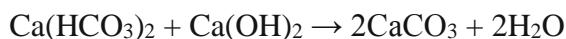
Costly equipment and costly chemicals are needed for regeneration.

Turbid water can not be used as it decreases the efficiency of resins.

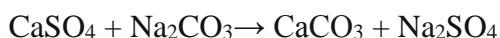
It can be operated only small scale purification of water.

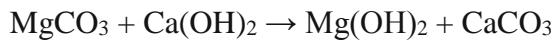
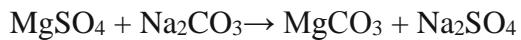
Soda lime is a process used in water treatment to remove **Hardness** from water. This process is now obsolete but was very useful for the treatment of large volumes of hard water. Addition of lime (CaO) and soda (Na_2CO_3) to the hard water precipitates calcium as the carbonate, and magnesium as its hydroxide. The amounts of the two chemicals required are easily calculated from the analysis of the water and stoichiometry of the reactions. The lime-soda uses lime, $Ca(OH)_2$ and soda ash, Na_2CO_3 , to precipitate hardness from solution. Carbon dioxide and carbonate hardness (calcium and Magnesium bicarbonate) are complexed by lime. In this process Calcium and Magnesium ions are precipitated by the addition of lime ($Ca(OH)_2$) and soda ash (Na_2CO_3).

temporary water hardness the following reactions are take places:



In case of permanent water hardness the following reactions are take places:





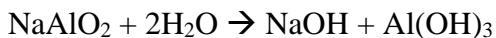
There are two type lime-soda process; cold lime soda process and hot lime soda process.

Cold Lime soda process

When the chemicals (L Or S) are added to hard water at room temperature, the process is known as cold lime soda process.

At room temperature, the precipitates are finely divided and do not settle easily, nor can they be easily filtered.

They help in the formation of coarse precipitates.



Cold lime soda process provides water containing a residual hardness of 50-60 ppm.

Hot Lime soda process

When the chemicals are added to higher temperature (80°C to 150°C) the process is known as Hot lime soda process.

Why Hot L-S Method is proffered over the Cold L-S Method

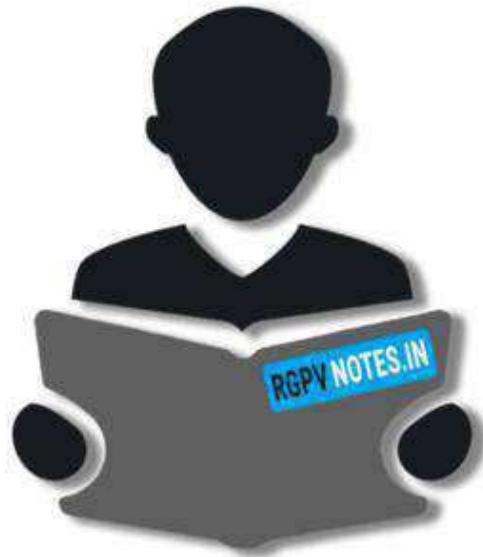
- In hot lime soda method at higher temperature.
- The reactions are fast.
- Precipitation is more complete.
- Settling rate and filtration rates are increased. As a result the precipitate and sludge formed settle rapidly and coagulants and also chemicals needed are in smaller quantities than that of cold L-S process.
- Dissolved gases are eliminated from water to a certain extent.
- The softened water recovered has hardness of about 15 to 30 ppm.

Advantages

- It is very economical process.
- Less amount of coagulants are required.
- Certain quantity of minerals is reduced from water.
- The level of carbon dioxide can also be reduced using hot L-S Process.
- Fe and Mn are also removed from the water.
- The process increases the PH value of the treated water thereby corrosion of the distribution pipes is reduced.

Disdvantages

- Sludge disposal is problem.
- This can remove hardness only up to 15 ppm, which is not good for boilers.
- Careful operations and skilled supervision are required for economical and efficient softening.



RGPVNOTES.IN

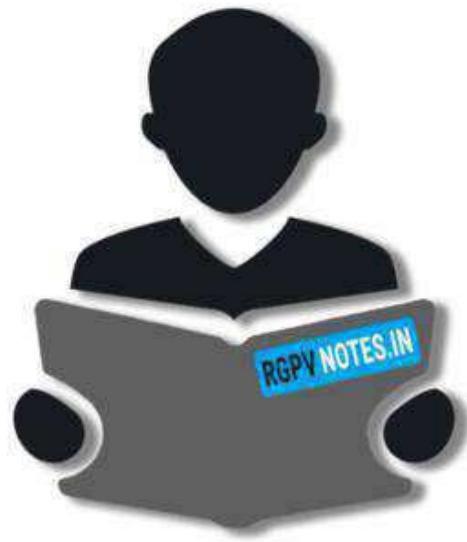
We hope you find these notes useful.

You can get previous year question papers at
<https://qp.rgpvnotes.in> .

If you have any queries or you want to submit your
study notes please write us at
rgpvnotes.in@gmail.com



LIKE & FOLLOW US ON FACEBOOK
facebook.com/rgpvnotes.in



RGPVNOTES.IN

Program : **B.Tech**

Subject Name: **Engineering Graphics**

Subject Code: **BT-105**

Semester: **1st**



LIKE & FOLLOW US ON FACEBOOK
facebook.com/rgpvnotes.in

Module – II

Orthographic Projections covering, Principles of Orthographic Projections - Conventions - Projections of Points and lines inclined to both planes; Projections of planes inclined Planes – Auxiliary Planes

Projection:

Any object has three dimensions, viz., length, width and thickness. A projection is defined as a representation of an object on a two dimensional plane. The projections of an object should convey all the three dimensions, along with other details of the object on a sheet of paper.

Types of Projections

1. Pictorial projections
 - (i) Perspective projection
 - (ii) Isometric projection
 - (iii) Oblique projection
2. Orthographic Projections

1. Pictorial Projections

The Projections in which the description of the object is completely understood in one view is known as pictorial projection. They have the advantage of conveying an immediate impression of the general shape and details of the object, but not its true dimensions or sizes.

2. Orthographic Projection

'ORTHO' means right angle and orthographic means right angled drawing. When the projectors are perpendicular to the plane on which the projection is obtained, it is known as orthographic projection.

Method of Obtaining Front View

Imagine an observer looking at the object from an infinite distance (Fig. 2.1). The rays are parallel to each other and perpendicular to both the front surface of the object and the plane. When the observer is at a finite distance from the object, the rays converge to the eye as in the case of perspective projection. When the observer looks from the front surface F or the block, its true shape and size is seen. When the rays or projectors are extended further they meet the vertical plane (V.P) located behind the object. By joining the projectors meeting the plane in correct sequence the Front view (Fig. 2.1) is obtained.

Front view shows only two dimensions of the object, Viz. length L and height H. It does not show the breadth B. Thus one view or projection is insufficient for the complete description of the object.

As Front view alone is insufficient for the complete description of the object, another plane called Horizontal plane (H.P) is assumed such that it is hinged and perpendicular to Y.P and the object is in front of the Y.P and above the H.P as shown in Fig. 2.1.

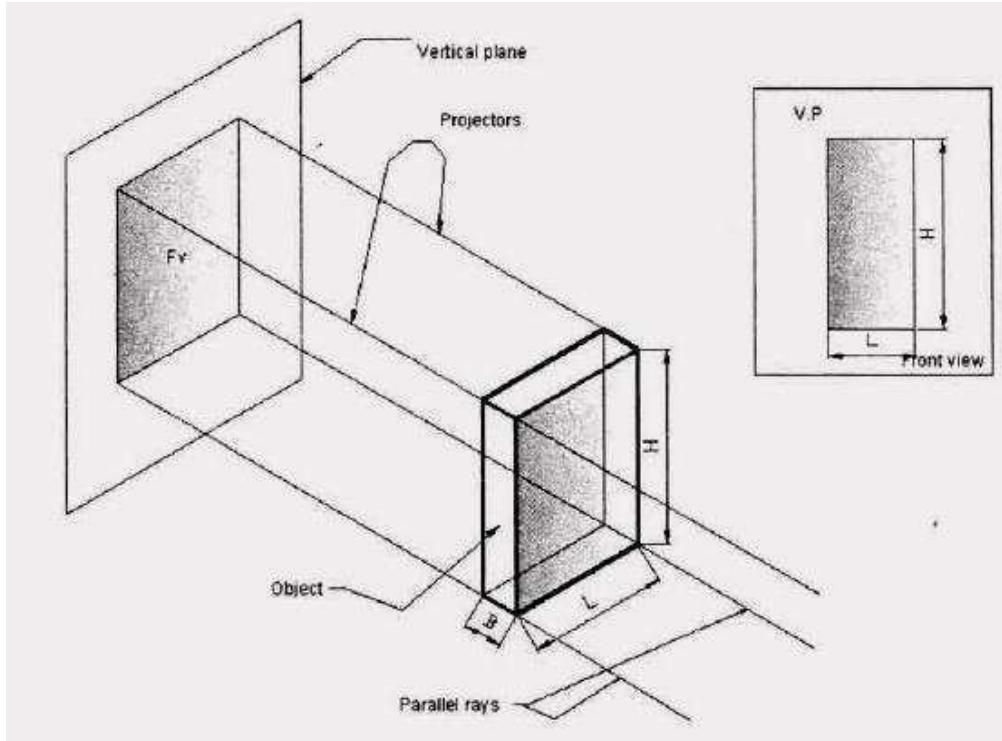


Fig. 2.1 Method of Obtaining Orthographic Front View

Method of Obtaining Top View

Looking from the top, the projection of the top surface is the Top view (TV). Both top surface and Top view are of exactly the same shape and size. Thus, Top view gives the True length L and breadth B of the block but not the height H.

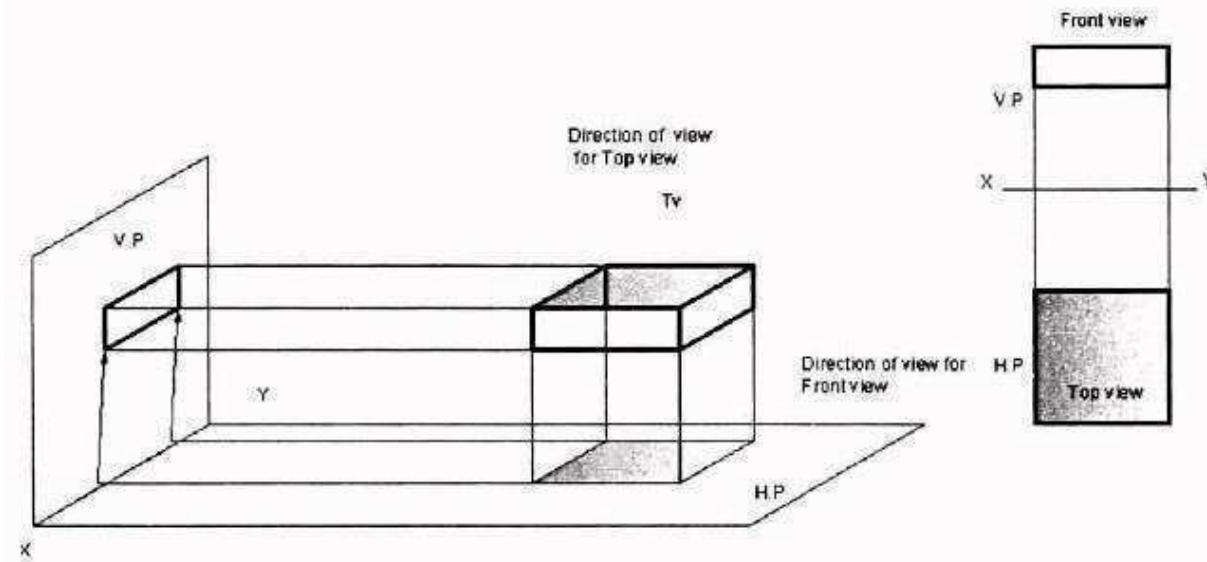


Fig. 2.2 Method of Obtaining Orthographic Top View

XY Line: The line of intersection of VP and H.P is called the reference line and is denoted as XY.

Obtaining the Projection on the Drawing Sheet

It is convention to rotate the H.P through 900 in the clockwise direction about XY line so that it lies in the extension of V.P. as shown in Fig. 2.2 a. The two projections Front view and Top view may be drawn on the two dimensional drawing sheet as shown in Fig. 2.2 b.

Thus, all details regarding the shape and size, Viz. Length (L), Height (H) and Breadth (B) of any object may be represented by means of orthographic projections i.e., Front view and Top view.

Terms Used

VP and H.P are called as Principal planes of projection or reference planes. They are always transparent and at right angles to each other. The projection on VP is designated as Front view and the projection on H.P as Top view.

Four Quadrants

When the planes of projections are extended beyond their line of intersection, they form Four Quadrants. These quadrants are numbered as I, II, III and IV in clockwise direction when rotated about reference line XY as shown in Fig. 2.3.

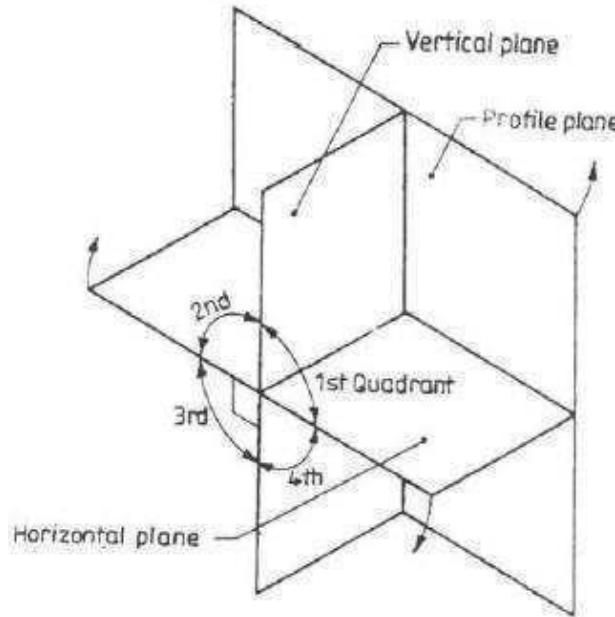


Fig. 2.3 Four Quadrants

In the Figure 2.4 the object is in the first quadrant and the projections obtained are "First angle projections" i.e., the object lies in between the observer and the planes of projection. Front view shows the length (L) and height (H) of the object, and Top view shows the length (L) and the breadth (B) of it.

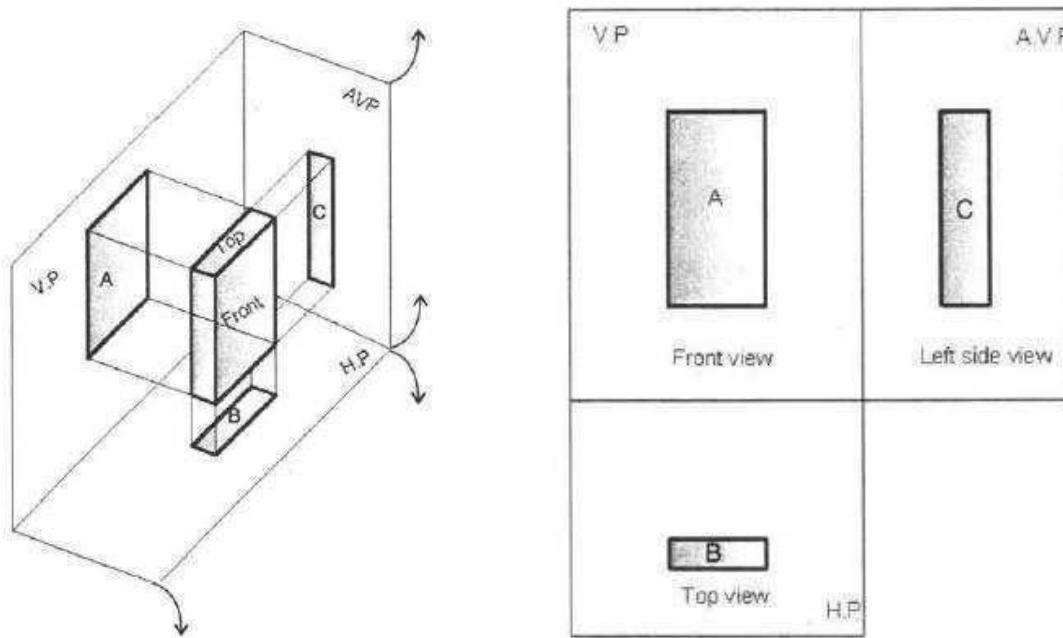


Fig. 2.4 Orthographic Projection of Front, Top and Side views

The object may be situated in anyone of four quadrants, its position relative to the planes being described as in front of V.P. and above H.P. in the first quadrant and so on.

Figure shows the two principle planes H.P and V.P. and another Auxiliary vertical plane (AVP). AVP is perpendicular to both V.P. and H.P. Front view is drawn by projecting the object on the V.P. Top view is drawn by projecting the object on the H.P. The projection on the AVP as seen from the left of the object and drawn on the right of the front view is called left side view.

First Angle Projection

When the object is situated in First Quadrant, that is, in front of V.P. and above H.P., the projections obtained on these planes are called First angle projection.

- (i) The object lies in between the observer and the plane of projection.
- (ii) The front view is drawn above the XY line and the top view below XY. (Above XY line is V.P. and below XY line is H.P.).
- (iii) In the front view, H.P coincides with XY line and in top view V.P. coincides with XY line.
- (iv) Front view shows the length (L) and height (H) of the object and Top view shows the length (L) and breadth (B) or width (W) or thickness (T) of it.

Third Angle Projection

In this, the object is situated in Third Quadrant. The Planes of projection lie between the object and the observer. The front view comes below the XY line and the top view about it.

Projection of Points

A solid consists of number of planes, a plane consists of a number of lines and a line in turn consists of number of points, from this, it is obvious that a solid may be generated by a plane.

Points in Space

A point may lie in space in anyone of the four quadrants. The positions of a point are:

1. First quadrant, when it lies above Horizontal Plane and in front of Vertical Plane.
2. Second quadrant, when it lies above Horizontal Plane and behind Vertical Plane.
3. Third quadrant, when it lies below Horizontal Plane and behind Vertical Plane.
4. Fourth quadrant, when it lies below Horizontal Plane and in front of Vertical Plane.

Convention

- Top views are represented by only small letters e.g. a.
- Their front views are conventionally represented by small letters with dashes e.g. a'
- Profile or side views are represented by small letters with double dashes e.g. a''
- The line of intersection of HP and VP is denoted as XY.
- The line of intersection of VP and PP is denoted as X₁Y₁
- Projectors and the lines of the intersection of planes of projections are shown as thin lines.

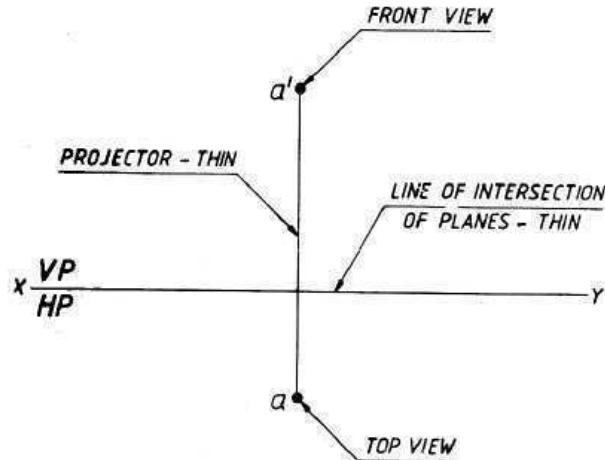


Fig. 2.5

Point in the First quadrant

Example: Point A is 40 mm above H.P. and 60 mm in front of V.P. Draw its front and top view.

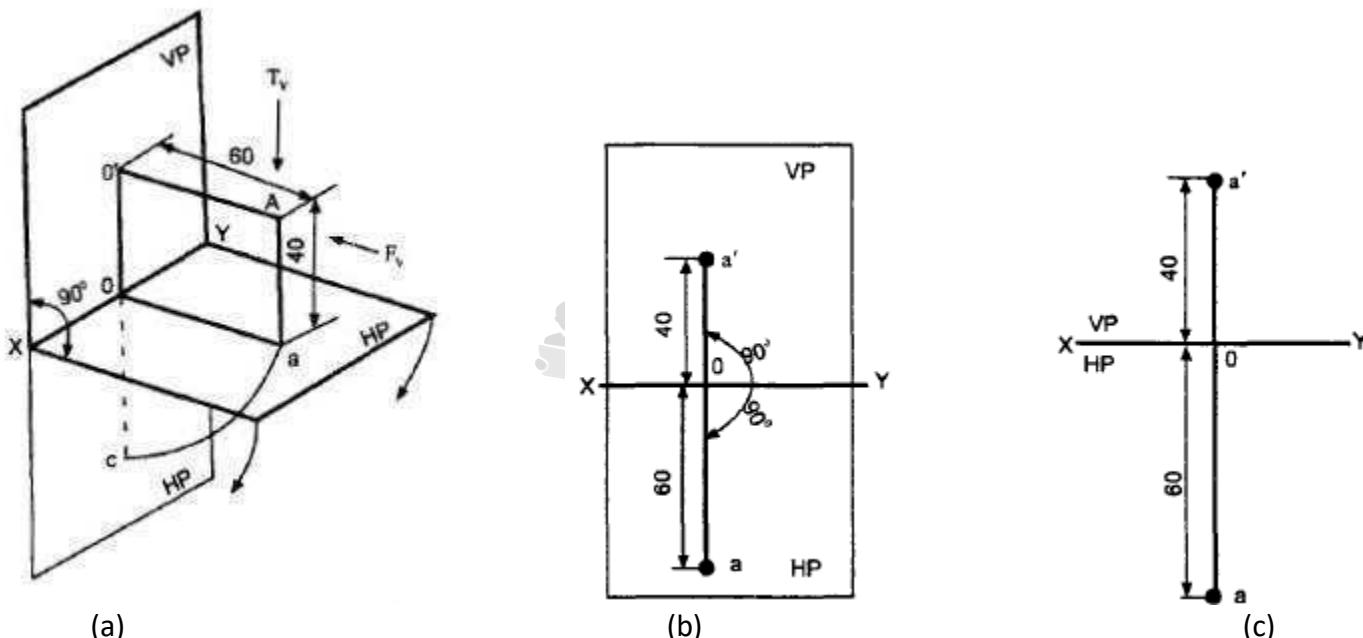


Fig. 2.6

- Draw a thin horizontal line, XY, to represent the line of intersection of HP and VP.
- Draw the Top View (a) 60 mm below the XY line.
- Draw the projector line starts from a in vertical direction.
- Draw the Front View (a') 40 mm above the XY line.

Projection of Lines

The shortest distance between two points is called a straight line. The projectors of a straight line are drawn therefore by joining the projections of its end points. The possible projections of straight lines with respect to V.P and H.P in the first quadrant are as follows:

1. Perpendicular to one plane and parallel to the other.
2. Parallel to both the planes.
3. Parallel to one plane and inclined to the other.
4. Inclined to both the planes.

1. Line perpendicular to H.P and parallel to V.P

The pictorial view of a straight line AB in the First Quadrant is shown in Fig. 2.7 (a).

1. Looking from the front; the front view of AB, which is parallel to V.P. and marked, $a' b'$, is obtained. True length of AB = $a' b'$

2. Looking from the top; the top view of AB, which is perpendicular to H.P is obtained a and b coincide.

3. The Position of the line AB and its projections on H.P. and V.P. are shown in Fig. 2.7 (b).

4. The H.P is rotated through 90° in clock wise direction as shown in Fig. 2.7 (b).

5. The projection of the line on V.P which is the front view and the projection on H.P, the top view are shown in Fig. 2.7 (c).

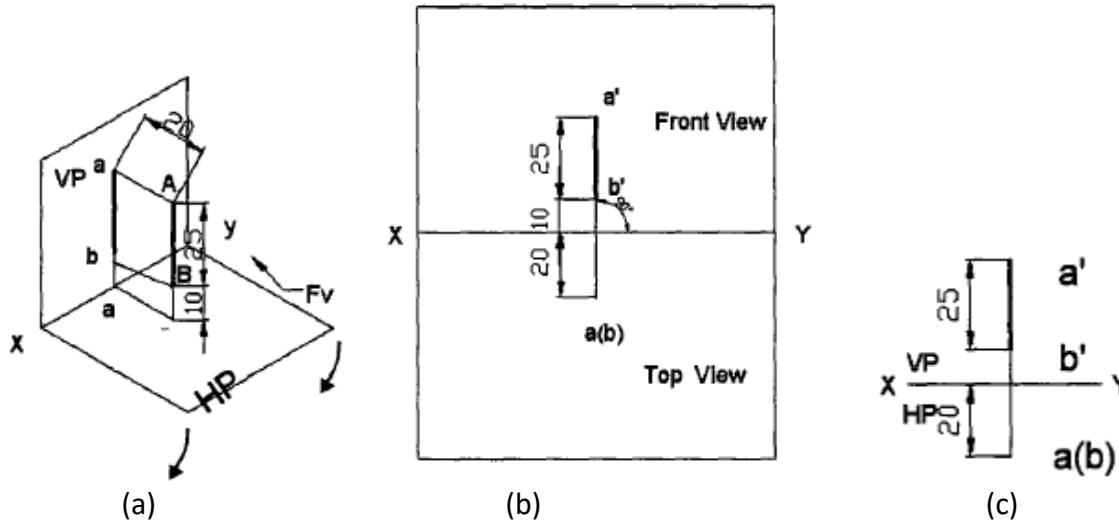


Fig. 2.7 Line perpendicular to H.P and parallel to V.P

2. Line perpendicular to V.P. and parallel to H.P.

The line is parallel to H.P. Therefore the true length of the line is seen in the top view. So, top view is drawn first.

1. Draw XY line and draw a projector at any point on it.

2. Point A is 20 mm in front of V.P. Mark a which is the top view of A at a distance of 20 mm below XY on the projector.

3. Mark the point b on the same projector at a distance of 50 mm below a. ab is the top view which is true length of AB.

4. To obtain the front view; mark $b'(a')$ at a distance 40mm above XY line on the same projector.

5. The line AB is perpendicular to V.P. So, the front view of the line will be a point. Point A is hidden by B. Hence the front view is marked as $b'(a')$. b' coincides with a' .

6. The final projections are shown in Fig. 2.8 (c).

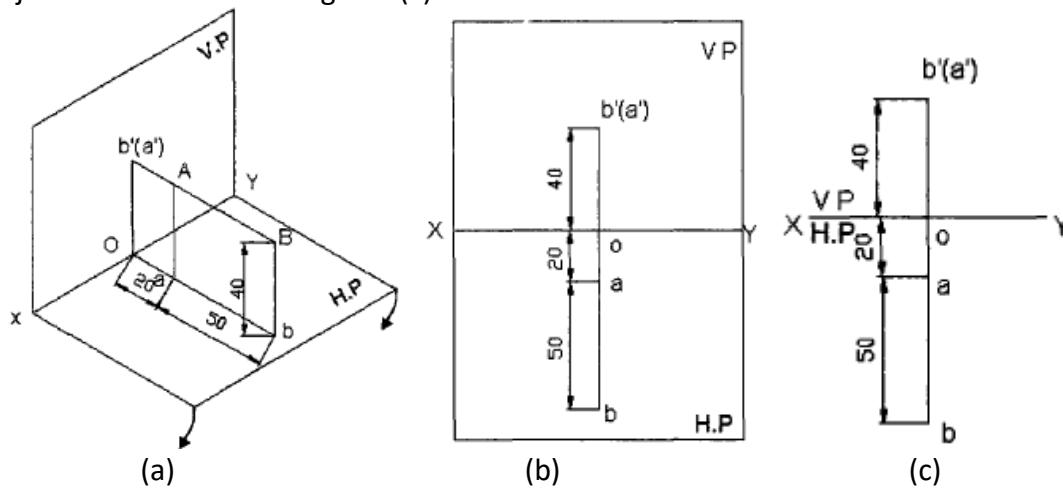


Fig. 2.8 Line perpendicular to V.P. and parallel to H.P.

3. Line parallel to both the planes

1. Draw the XY line and draw a projector at any point on it.
2. To obtain the front view mark c' at a distance of 40 mm above XY (H.P.). The line CD is parallel to both the planes. Front view is true length and is parallel to XY. Draw $c'd'$ parallel to XY such that $c'd' = CD = 30 \text{ mm}$, which is the true length.
3. To obtain the top view; the line is also parallel to V.P. and 20 mm in front of V.P. Therefore on the projector from c' , mark c at distance 20 mm below XY line.
4. Top view is also true length and parallel to XY. Hence, cd parallel to XY such that $cd = CD = 30 \text{ mm}$ is the true length (Fig. 2.9).

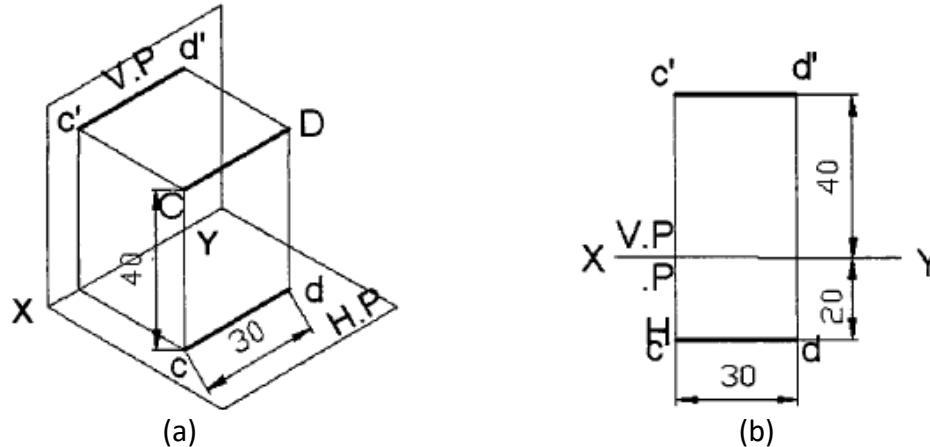


Fig. 2.9 Line parallel to both the planes

4. Line parallel to V.P. and inclined to H.P.

1. A is 15 mm above H.P mark a' , 15 mm above XY.
2. A is 20 mm in front of V.P. Hence mark a 20 mm below XY.
3. To obtain the front view $a'b'$; as AB is parallel to V.P and inclined at an angle α to H.P., $a'b'$ will be equal to its true length and inclined at an angle of 30° to H.P. Therefore draw a line from a' at an angle 30° to XY and mark b' such that $a'b' = 40 \text{ mm} = \text{true length}$.
4. To obtain the top view ab ; since the line is inclined to H.P its projection on H.P (its top view) is reduced in length. From b' draw a projector to intersect the horizontal line drawn from a at b . ab is the top view of AB. Inclination of line with the H.P is always denoted as θ .

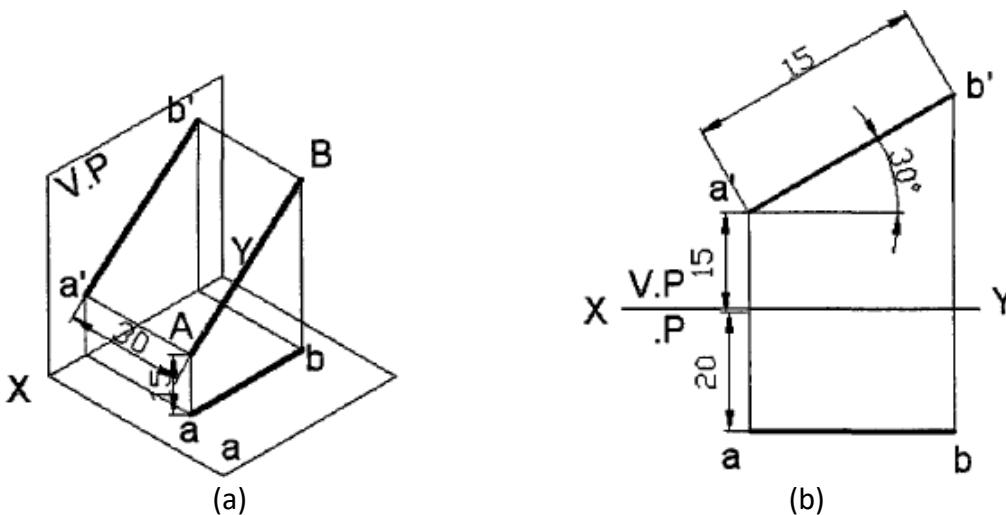


Fig. 2.10 Line parallel to V.P. and inclined to H.P.

5. Line parallel to H.P. and inclined to V.P.

1. A is 30 mm above H.P, mark a' , 30 mm above XY.
 2. A is 20 mm in front of V.P., mark a 20 mm below XY.
 3. To obtain the top view; as AB is parallel to H.P and inclined at an angle Φ to V.P, ab will be equal to the true length of AB, and inclined at angle Φ to XY. Therefore, draw a line from a at 40° to XY and mark b such that $ab = 60$ mm true length.
 4. To obtain the front view $a'b'$, since the line is inclined to V.P its projection on V.P. i.e., the front view will be reduced in length. Draw from b a projector to intersect the horizontal line drawn from a at bl . $a'b'$ is the front view of AB.
- Inclination of a line with V.P is always denoted by Φ .

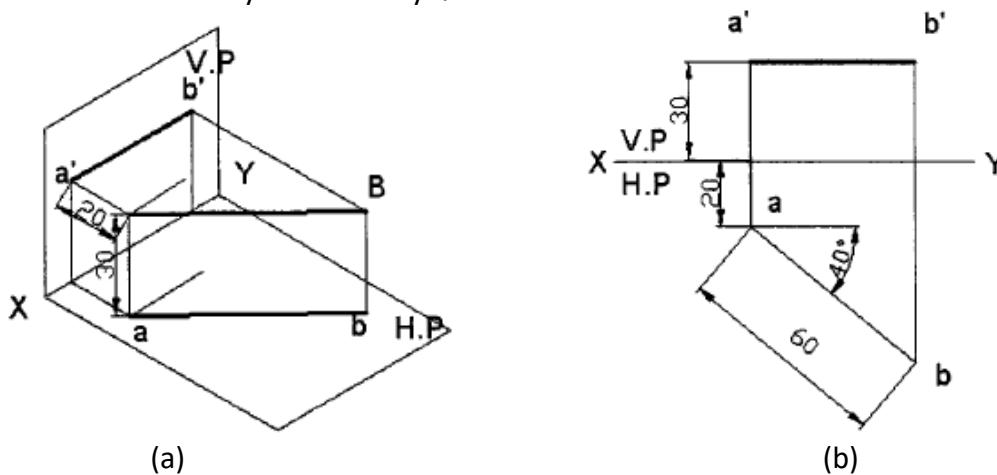


Fig. 2.11 Line parallel to H.P. and inclined to V.P.

6. Line inclined to both the planes

When a line is inclined to both H.P and V.P, it is called an oblique line. The solution to this kind of problem is obtained in three stages, as described below.

Stage I Assume the line is inclined to H.P by θ and parallel to V.P. (Fig. 2.12 b)

1. Draw the projections $a'b'_1$ and ab_1 of the line AB_1 ($=AB$), after locating projections and a from the given position of the end A.

Keeping the inclination θ constant rotate the line AB_1 to AB, till it is inclined at Φ° to V.P. This rotation does not change the length of the top view ab^l and the distance of the point B_1 ($=B$) from H.P. Hence, (i) the length of ab^l is the final length of the top view and (ii) the line f-f, parallel to XY and passing through b^l is the locus of the front view of the end of point B.

Stage II Assume the line is inclined to VP by Φ° and parallel to H.P (Fig. 2.12 c)

2. Draw the projections ab_2 and ab of the line AB_2 ($=AB$), after locating the projections a' and a, from the given position of the end A.

Extending the discussion on the preceding stage to the present one, the following may be concluded. (i) The length ab is the final length of the front view and (ii) the line t-t, parallel to XY and passing through b_2 is the locus of the top view of the end point B.

Stage III Combine Stage I and Stage II (Fig. 2.12 d),

3. Obtain the final projections by combining the results from stage I and II as indicated below:

- (i) Draw the projections $a'b'_1$ and ab_2 making an angle θ and Φ respectively with XY, after location of the projections a' and a, from the given position of the end point A.
- (ii) Obtain the projections $a'b'_1$ and ab^l , parallel to XY, by rotation.
- (iii) Draw the lines f-f and t-t the loci parallel to XY and passing through b^l and b_2 respectively.
- (iv) With centre a' and radius $a'b'_1$, draw an arc meeting f-f at b^l .
- (v) With centre a and radius ab_2 , draw an arc meeting t-t at b.
- (vi) Join a' , b^l and a, b forming the required final projections. It is observed from the fig. 2.12 c that:

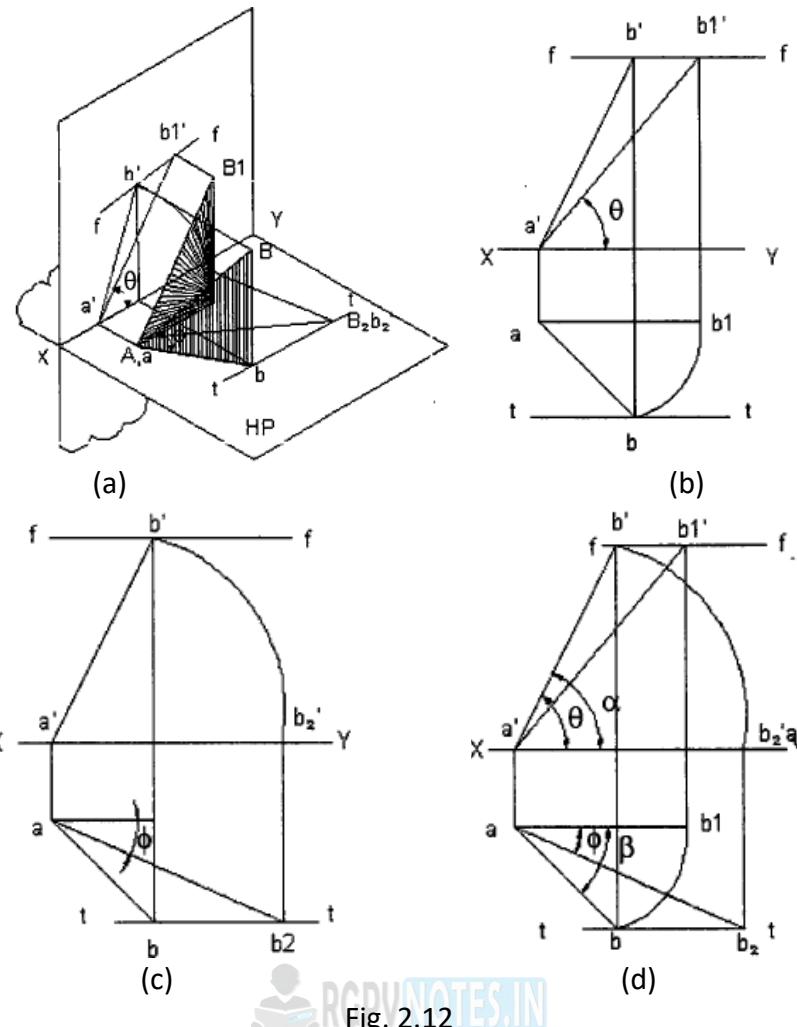


Fig. 2.12

To determine the true length of a line, given its projections - Rotating line method

In this, each view is made parallel to the reference line and the other view is projected from it. This is exactly reversal of the procedure adopted in the preceding construction.

Construction:

1. Draw the given projections $a'b'$ and ab .
2. Draw $f-f$ and $t-t$, the loci passing through b' and b and parallel to XY .
3. Rotate $a'b'$ to $a'b_2'$, parallel to XY .
4. Draw a projector through b_2' to meet the line $t-t$ at b_2 .
5. Rotate ab_1 parallel to XY .
6. Draw a projector through b_1 to meet the line $f-f$ at b_1'
7. Join a' , b_1' and a , b_2 .
8. Measure and mark the angles θ and Φ .

The length $a'b_1'$ ($= ab_2$) is the true length of the given line and the angles θ and Φ , the true inclinations of the line with H.P and V.P. respectively.

Projection of Planes

A plane figure has two dimensions viz. the length and breadth. It may be of any shape such as triangular, square, pentagonal, hexagonal, circular etc. The possible orientations of the planes with respect to the principal planes H.P. and V.P. of projection are:

1. Plane parallel to one of the principal planes and perpendicular to the other,
2. Plane perpendicular to both the principal planes,
3. Plane inclined to one of the principal planes and perpendicular to the other,
4. Plane inclined to both the principal planes.

1. Plane parallel to one of the principal planes and perpendicular to the other:

When a plane is parallel to V.P the front view shows the true shape of the plane. The top view appears as a line parallel to XY. Fig. 2.13 (a) shows the projections of a square plane ABCD, when it is parallel to V.P. and perpendicular to H.P. The distances of one of the edges above H.P. and from the V.P. are denoted by d_1 and d_2 respectively.

Fig. 2.13 (b) shows the projections of the plane. Fig. 2.13 (c) shows the projections of the plane, when its edges are equally inclined to H.P.

Fig. 2.13 (d) shows the projections of a circular plane, parallel to H.P and perpendicular to V.P.

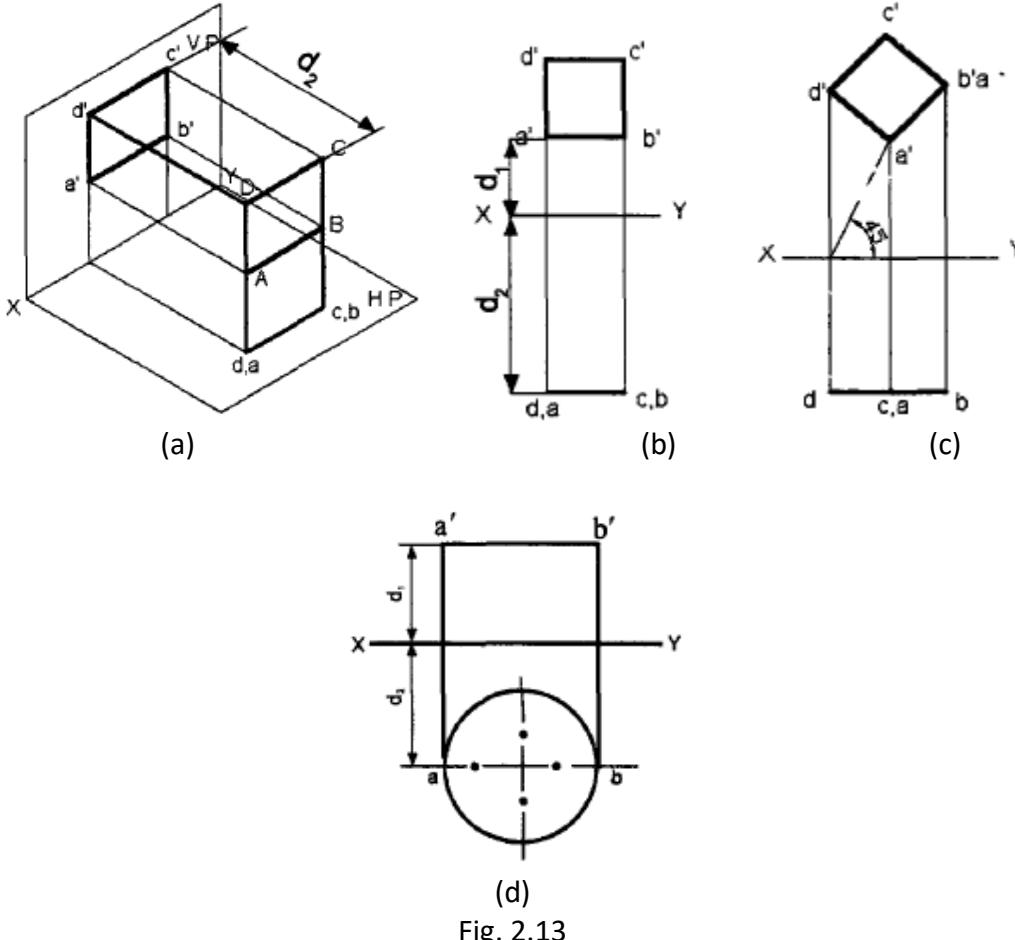


Fig. 2.13

2. Plane perpendicular to both H.P and V.P.

When a plane is perpendicular to both H.P. and V.P, the projections of the plane appear as straight lines. Fig. 2.14 shows the projections of a rectangular plane ABCD, when one of its longer edges is parallel to H.P. Here, the lengths of the front and top views are equal to the true lengths of the edges.

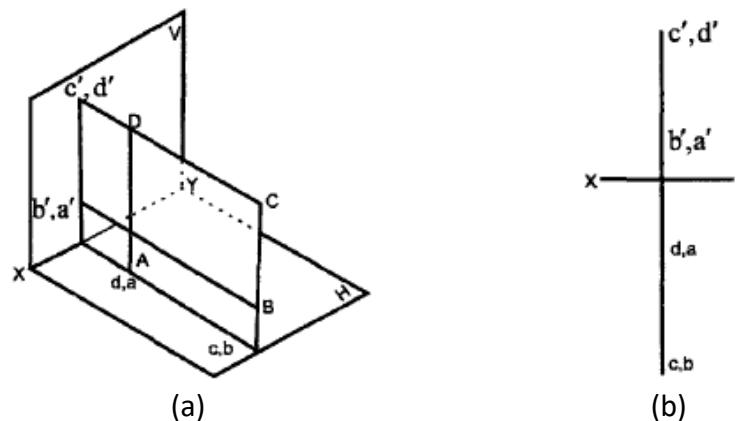


Fig. 2.14

3. Plane inclined to one of the principal planes and perpendicular to the other

a) Plane inclined to H.P. and perpendicular to V.P.:

When a plane is inclined to one plane and perpendicular to the other, the projections are obtained in two stages.

Problem: Projections of a pentagonal plane ABCDE, inclined at θ to H.P and perpendicular to V.P. and resting on one of its edges on H.P.

Construction:

Stage I Assume the plane is parallel to H.P (lying on H.P) and perpendicular to V.P.

1. Draw the projections of the pentagon ABCDE, assuming the edge AE perpendicular to V.P. $a'e'$ $b_1'd_1'c_1'$ on XY is the front view and $ab_1c_1d_1e$ is the top view.

Stage II Rotate the plane (front view) till it makes the given angle with H.P.

2. Rotate the front view till it makes the given angle θ with XY which is the final front view.
3. Obtain the final top view abcde by projection.

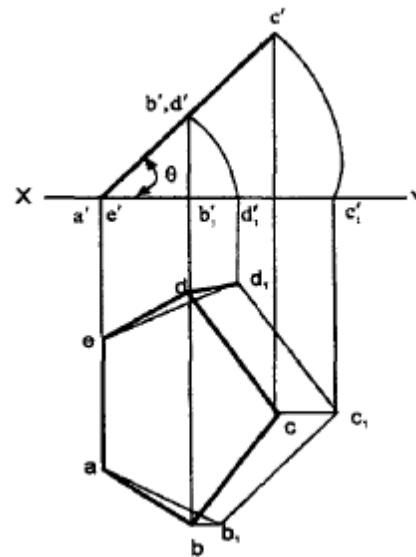


Fig. 2.15 Plane inclined to H.P. and perpendicular to V.P.

b) Plane inclined to V.P. and perpendicular to H.P.:

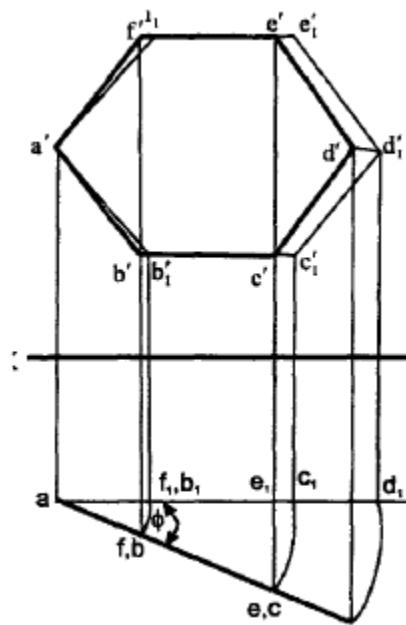


Fig. 2.16 Plane inclined to V.P. and perpendicular to H.P.

4. Plane inclined to both H.P and V.P

If a plane is inclined to both H.P and V.P, it is said to be an oblique plane. Projections of oblique planes are obtained in three stages.

Construction:

Stage I: Assume the plane is parallel to H.P and a shorter edge of it is perpendicular to V.P.

1. Draw the projections of the plane.

Stage II: Rotate the plane till it makes the given angle with H.P.

2. Redraw the front view, making given angle θ with XY and then project the top view.

Stage III: Rotate the plane till its shorter edge makes the given angle Φ with V.P.

3. Redraw the top view abed such that the shorter edge ad is inclined to XY by Φ .

4. Obtain the final front view $a_1b_1c_1d_1$, by projection.

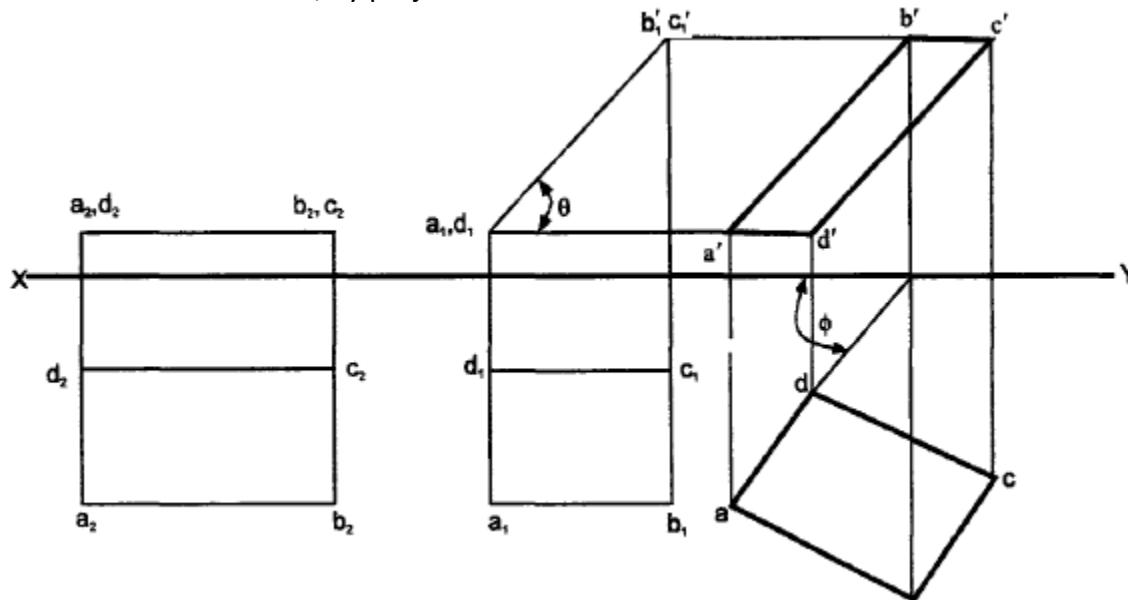
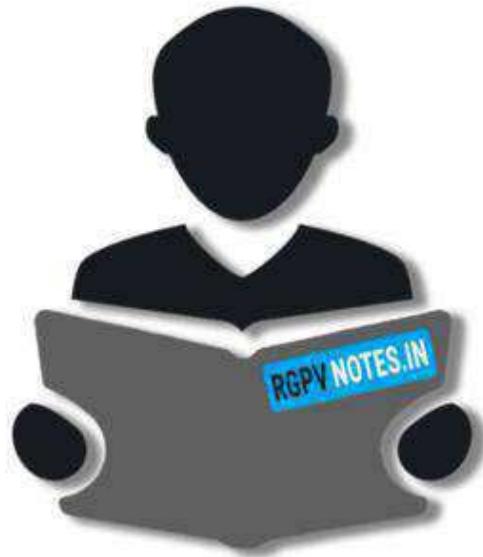


Fig. 2.17 Plane inclined to both the planes



RGPVNOTES.IN

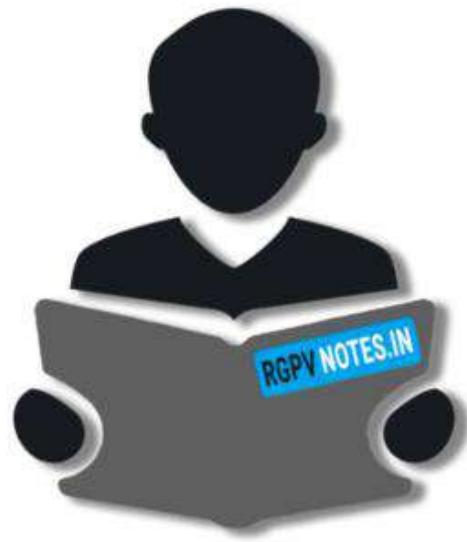
We hope you find these notes useful.

You can get previous year question papers at
<https://qp.rgpvnotes.in> .

If you have any queries or you want to submit your
study notes please write us at
rgpvnotes.in@gmail.com



LIKE & FOLLOW US ON FACEBOOK
facebook.com/rgpvnotes.in



RGPVNOTES.IN

Program : **B.Tech**

Subject Name: **Basic Electrical & Electronics Engineering**

Subject Code: **BT-104**

Year: **1st**



LIKE & FOLLOW US ON FACEBOOK
facebook.com/rgpvnotes.in

BASIC ELECTRICAL & ELECTRONICS ENGINEERING NOTES

Unit - III Magnetic Circuits - Syllabus

Basic definitions, Magnetization characteristics of Ferro magnetic materials, Magnetic field produced by current carrying conductor, Force on a current carrying conductor, AC excitation in magnetic circuits. Laws of electromagnetic Induction, induced voltage, direction of induced E.M.F, selfinductance and mutual inductance, energy in linear magnetic systems, Coils connected in series.

MAGNETIC CIRCUITS

Magnetic flux: The total number of flux lines coming out of the N-pole of a magnet is called the Magnetic flux. It is represented by Φ and the unit is weber

Magnetic flux density: It is the number of flux lines passing through an unit area of cross section held perpendicular to the lines of flux. It is given by the expression $B = \Phi/A$ its unit is Wb/m^2

Magnetic field strength or Magnetic intensity or Magnetizing force: The magnetic field strength at any point within a magnetic field is given by the force experienced by a unit N-pole of one weber placed at that point. It is represented by H and the unit is Newton/weber

Absolute permeability: A magnetic material when placed in a magnetic field acquires magnetization due to induction. A measure of the degree to which the lines of force of the magnetizing field can penetrate the medium is called the absolute permeability of the medium.

It is also defined as the ratio of flux density in that material to the magnetizing force producing that flux density. It is given by $\mu = B/H$ and the unit is Henry/metre

It is also given by $\mu = \mu_0\mu_r$ where $\mu_0 = 4\pi \times 10^{-7}$ Henry/metre

Relative permeability: It is given by the ratio of the flux density produced in that material to the flux density produced in vacuum by the same magnetizing force. It is given by $\mu_r = B/B_0$

Magneto motive force: It drives or tends to drive flux through a magnetic circuit or it is also equal to the work done in joules in carrying a unit magnetic pole once through the entire magnetic circuit. Its unit is Ampere turns.

Reluctance: It is defined as the property of a material that opposes the creation of magnetic flux in it. It is a measure of the opposition offered to the passage of magnetic flux through a material. It is given by $S = l/\mu_0\mu_r A$ or $S = mmf/\Phi$ its unit is AT/Wb.

Permeance - It is the reciprocal of reluctance. It is defined as the property of a material that initiates the development of magnetic flux. It is given by, $\text{Permeance} = \Phi/mmf$ its unit is Wb/AT or Henry.

Relations in magnetic circuits:

$$\text{We have, } H = \frac{NI}{l} \text{ or } NI = Hl ; \quad B = \mu_0\mu_r H \text{ or } H = \frac{B}{\mu_0\mu_r} ; \quad B = \frac{\Phi}{A} ; \quad NI = \frac{B}{\mu_0\mu_r} \times l$$

$$\text{Amp. turns for any material, } AT_M = \frac{B}{\mu_0\mu_r} \times l_M ; \quad \text{Amp. turns for airgap, } AT_G = \frac{B}{\mu_0} \times l_G$$

Leakage flux: Whenever a magnetic material is magnetized flux lines are produced in the material. The flux lines existing in the material and the air gap is called Useful flux Φ_U . The flux lines not existing in the

BASIC ELECTRICAL & ELECTRONICS ENGINEERING NOTES

material and existing around the coil wound on the magnetic material is called leakage flux Φ_L . Leakage flux is the flux that follows a path not intended for it. The total flux produced by the solenoid $\Phi = \Phi_U + \Phi_L$. The ratio of total flux produced by the solenoid to the useful flux set up in the material and air gap is known as leakage coefficient. It is given by $\lambda = \Phi/\Phi_U$ and its value is more than 1.

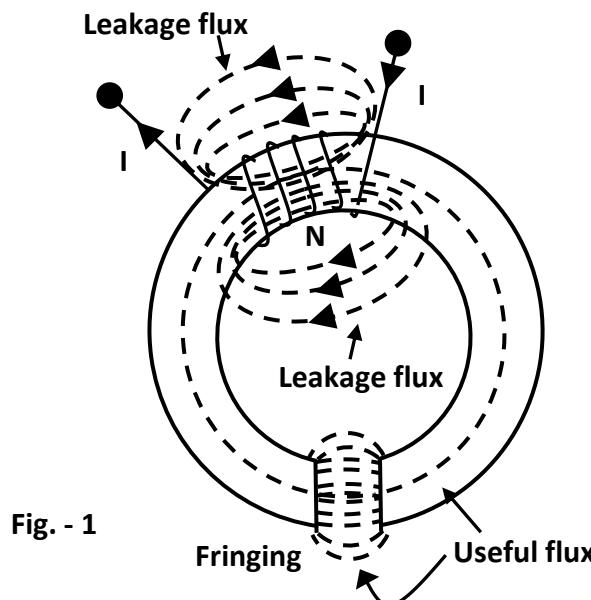


Fig. - 1

Fringing is the bulging of the magnetic flux lines in the air gap of a magnetic material. This increases the effective area in the air gap and decreases the flux density. Every magnetic material with air gaps suffer from the problem of fringing. Fringing is directly proportional to the length of the air gap. Fig.-1 shows a magnetic circuit with mmf (NI), Total flux, Useful flux and Leakage flux. The effect of Fringing is also shown in the air gap of the magnetic material.

Series magnetic circuit:

A magnetic circuit that is made up of a number of magnetic materials of different cross sectional areas, of different lengths, of different relative permeability along with an air gap carrying the same value of flux represents a series magnetic circuit. This magnetic circuit will have only one path for the magnetic flux.

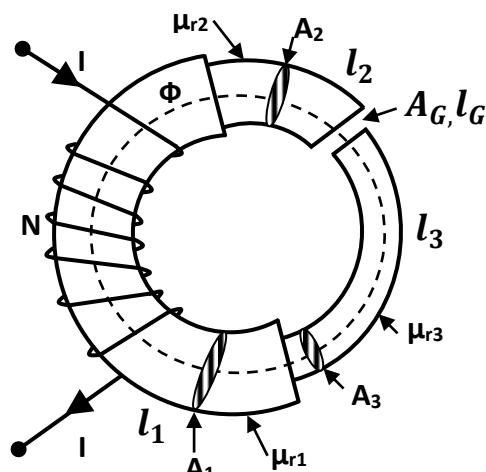


Fig. - 2

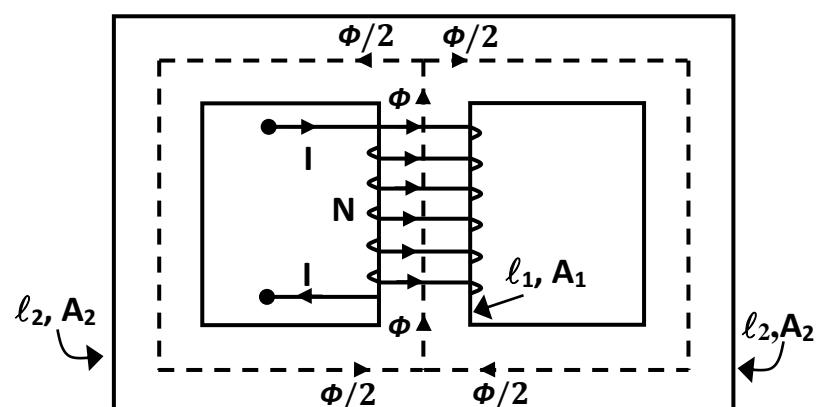


Fig. - 3

BASIC ELECTRICAL & ELECTRONICS ENGINEERING NOTES

A series magnetic circuit made of three parts is as shown in fig. -2. The total ampere turns required for the magnetic circuit will be equal to the sum of all the ampere turns required for each of the magnetic material as well as the air gap. Total AT = $AT_1 + AT_2 + AT_3 + AT_G$

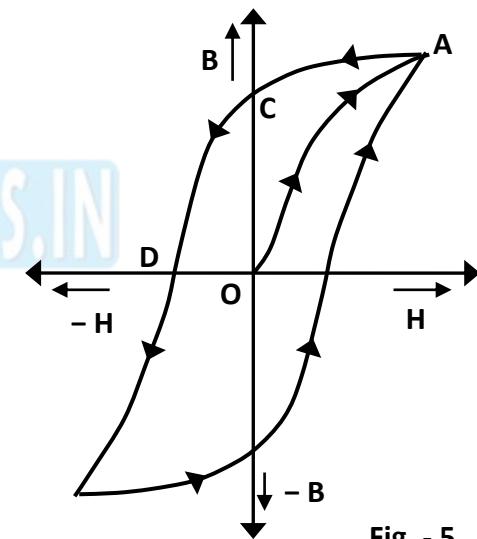
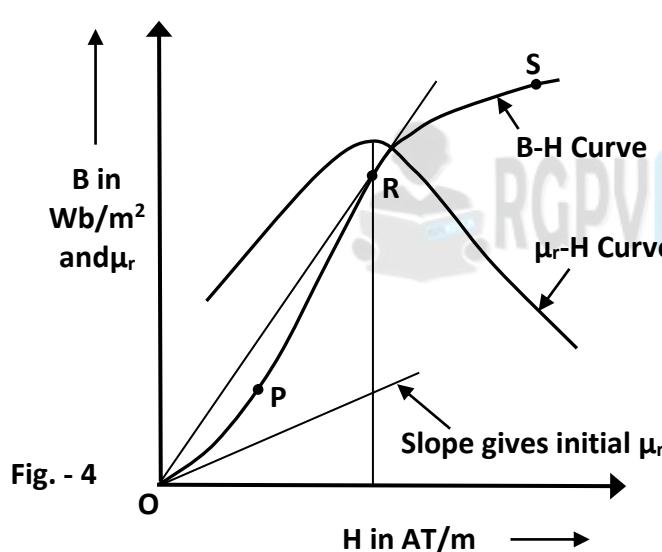
Parallel magnetic circuit:

A magnetic circuit that has two or more than two paths for the magnetic flux is called a parallel magnetic circuit. In these circuits the value of flux will not be the same in all the parts of the magnetic circuit. A parallel magnetic circuit can also have an air gap in the central limb or in the side limbs or in all the limbs. A simple parallel magnetic circuit is as shown in fig. - 3. The total ampere turns required for the magnetic circuit will be equal to the sum of the ampere turns required for the central limb and the ampere turns required for one of the side limbs. It indicates that the ampere turns required for one side limb is capable of driving flux in both the side limbs.

Total AT = Ampere turns required for central limb + Ampere turns required for any one side limb.

$$\text{Total AT} = AT_C + AT_S$$

Magnetization characteristics of Ferromagnetic materials:



When an iron specimen is subjected to an increasing value of magnetizing force, H and the corresponding values of flux density B plotted against H helps in realizing the B-H curve shown in fig.-4. Using the relation $\mu_r = B/\mu_0 H$ the curve of relative permeability μ_r against H may be obtained from the plotted magnetisation characteristics. From the $\mu_r - H$ curve it is observed that the value of μ_r varies considerably with the value of H that helps in ascertaining the value of operating flux density. The B-H curve obtained can be divided into three regions - initial region OP, middle region PR and the region beyond R. The curve near the origin is nearly a straight line through the origin and the slope yields the value of initial permeability. The value of H at which the curve begins to bend varies over a wide range from material to material. In the middle region, increase in a small value of H will have a very large increase in the value of B. The slope is the greatest and the maximum value of permeability is within this region. At point R the tangent to the curve passes through the origin, where μ_r has its maximum value. In the region beyond R,

BASIC ELECTRICAL & ELECTRONICS ENGINEERING NOTES

any further increase in the value of H will have a small change in the value of B. The curve indicates the extent of magnetization due to the increase in value of H. The curve beyond point S shows that any further increase in the value of H will not have any variation in the value of B indicating the magnetic saturation of the material.

Hysteresis: Whenever an iron bar is taken through one cycle of magnetization a Hysteresis loop shown in fig. - 5 is traced. The area of the loop represents the total energy spent in taking the iron bar through one cycle of magnetization. The Hysteresis loop is a plot of B and H, it is also defined as the lagging of flux density B behind the magnetizing force H. When H is increased initially the curve OA is traced and the iron bar gets magnetically saturated. As and when H is decreased B also decreases. But when H is zero the value of B is not zero. The distance OC on the Y-axis represents the Residual flux density B_r . The value of flux density B_r measures the retentivity of the material. To demagnetize the iron bar, we have to apply the magnetising force in the reverse direction. When H is reversed the value of B is reduced to zero at point D. The distance OD on the X-axis represents the Coercive force H_c . The value of H required to eliminate the residual magnetism is known as Coercive force and is a measure of the Coercivity of the material. Fig.-5 shows a Hysteresis loop traced for an iron bar taken through one cycle of magnetization.

Magnetic field produced by a current carrying conductor:

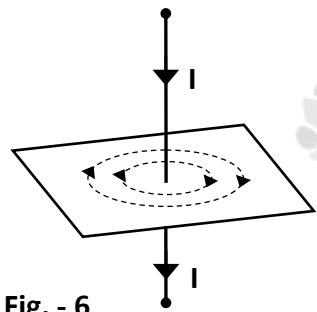


Fig. - 6

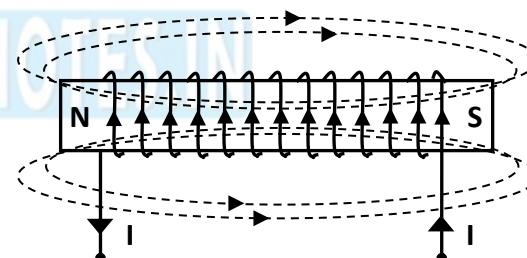


Fig. - 7

Let us consider a conductor passing through a card board as shown in fig.- 6, let a current I flow through it from top to the bottom. A magnetic compass when placed close to the wire has its needle pointing in a specific direction. The points of the needle are marked by moving the magnetic compass around the current carrying conductor. All the points when joined together appear in the form of a circle showing the existence of flux lines around the current carrying conductor. This simple experiment shows that a current carrying conductor will always have a magnetic field created around it. The Right hand thumb rule helps us to find out the direction of flux lines created around the current carrying conductor. If the current carrying conductor is grasped by the right hand such that the thumb held perpendicular to all the fingers points to the direction of current flow, then the fingers curled around the conductor points to the direction of magnetic flux around it.

Right hand thumb rule can also be applied to a coil wound around a magnetic material and carrying current as shown in fig.-7. If the coil is grasped by the right hand such that the four fingers curl around the coil

BASIC ELECTRICAL & ELECTRONICS ENGINEERING NOTES

pointing to the direction of current flow, the thumb held perpendicular to the four fingers points to the N-pole of the magnetic material.

Force on a current carrying conductor placed in a magnetic field:

Let us consider a current carrying conductor placed in a magnetic field of a permanent magnet. The flux lines produced by the magnet come out of the N pole and end up at the S pole. The flux lines of the magnetic field created by the current carrying conductor are in a circular form, aiding certain flux lines and opposing certain flux lines of the magnetic field of the magnet as shown in fig. - 8. The side on which the flux is aided becomes stronger and the side on which the flux gets opposed becomes weaker. The current carrying conductor is pushed from the side which has a stronger flux towards the side which has a weaker flux. Hence, due to the interaction between the two fluxes, a mechanical force is experienced by the current carrying conductor as shown in fig. - 9. The direction of the mechanical force experienced by the current carrying conductor placed in a magnetic field can be determined by Fleming's left hand rule.

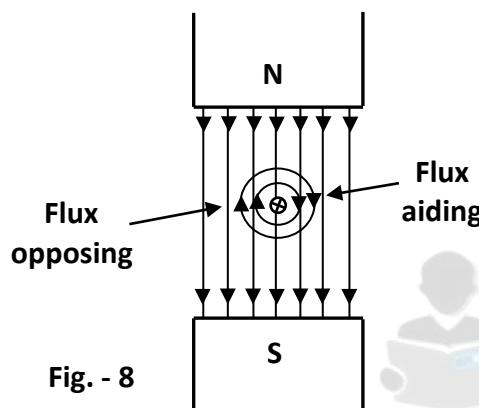


Fig. - 8

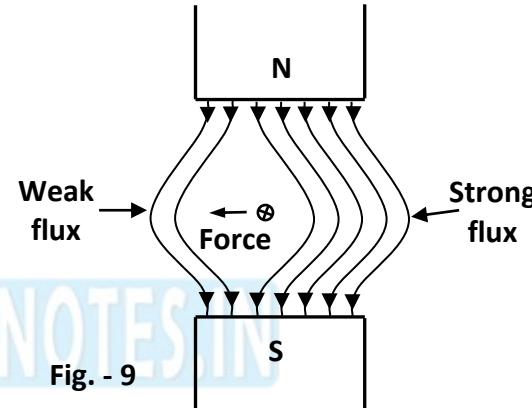


Fig. - 9

AC excitation of Magnetic circuit: A magnetic circuit with input given from an AC supply is called an AC magnetic circuit. Fig. - 10 shows a circular iron core wound with a coil of N turns. An alternating current, $i = I_m \sin \omega t$ flowing through the coil, creates an alternating flux Φ . As the current through the coil is sinusoidal in nature the flux created is also sinusoidal in nature varying with time.

The flux produced is given by $\Phi = \Phi_m \sin \omega t$

When the changing flux Φ links with N number of turns of the coil, emf is induced in the coil given by,

$$e = -N \frac{d\Phi}{dt} = -N \frac{d}{dt} (\Phi_m \sin \omega t)$$

$$= -N \omega \Phi_m \cos \omega t = N \omega \Phi_m (\sin \omega t - 90^\circ)$$

Maximum value of induced emf, $E_m = N \omega \Phi_m$

$$\therefore e = E_m (\sin \omega t - 90^\circ)$$

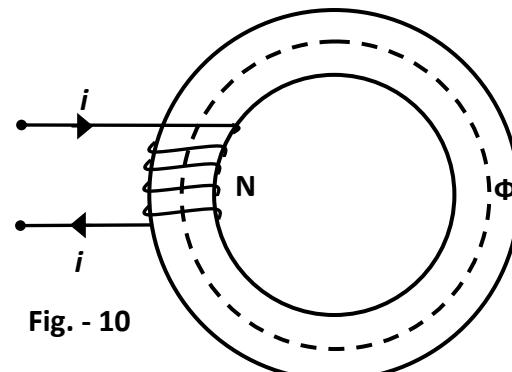


Fig. - 10

Electromagnetism: The phenomenon whereby an emf and hence current is induced in any conductor which is cut across or is cut by a magnetic flux is known as Electromagnetic induction.

BASIC ELECTRICAL & ELECTRONICS ENGINEERING NOTES

Faraday's laws of electromagnetic induction - First law: Whenever the flux linked with a circuit changes, an emf is always induced in it or whenever a conductor cuts across magnetic field lines, an emf is induced in that conductor.

Second law: The magnitude of the induced emf is equal to the rate of change of flux linkages.

$$e = -N \frac{d\Phi}{dt} \text{ volts}$$

The first law explains the phenomenon of Electromagnetic induction and the second law helps us to ascertain its magnitude.

Lenz's law: The direction of induced emf is always such that it tends to set up a current opposing the motion or the change of flux responsible for inducing that emf or The direction of induced emf is such as to oppose the very cause producing it.

Lenz's law helps us to ascertain the direction of the induced emf (statically)

Fleming's right hand rule: If the first finger of the right hand be pointed in the direction of the magnetic flux and if the thumb be pointed in the direction of motion of the conductor relative to the magnetic field, then the second finger held at right angles to both the thumb and the first finger, represents the direction of induced emf.

Fleming's right hand rule helps us to ascertain the direction of the dynamically induced emf

Fleming's left hand rule: If the first finger of the left hand be pointed in the direction of the magnetic flux and the second finger points in the direction of the current, then the thumb held at right angles to the first finger and second finger, would point in the direction of the mechanical force produced on the current carrying conductor.

Fleming's left hand rule helps us to find out the direction of the force experienced by a current carrying conductor when placed in a magnetic field.

Induced emf's: Electro magnetically induced emf's are of two types - Statically induced emf and Dynamically induced emf

Dynamically induced emf: Whenever a moving conductor cuts across magnetic flux lines, an emf is induced in the conductor called as dynamically induced emf or when a stationary conductor is cut by a rotating magnetic field, an emf is induced in the conductor.

This is observed in D.C and A.C. Generators.

If the conductor cuts the flux at right angles then the dynamically induced emf, $e = Blv$ volts and If the conductor moves at an angle θ with the direction of the lines of flux, then the dynamically induced emf, $e = Blv \sin\theta$ volts

Statically induced emf: Whenever a changing flux links with a stationary coil, an emf is induced in the coil called as statically induced emf. This is observed in Transformers.

$$\text{Statically induced emf, } e = -N \frac{d\Phi}{dt} \text{ volts}$$

BASIC ELECTRICAL & ELECTRONICS ENGINEERING NOTES

Statically induced emf's are of two types - Self induced emf and Mutually induced emf

Self induced emf: This is the emf induced in a coil due to the change of its own flux linked with it.

$$\text{We have self induced emf, } e_L = L \frac{dI}{dt} \text{ volts}$$

Mutually induced emf: This is the emf induced in one coil by the influence of another coil.

$$\text{We have mutually induced emf, } e_M = M \frac{dI_1}{dt} \text{ volts}$$

Self inductance: The property of a coil by virtue of which an emf is induced in it, whenever a changing current flows through it is called as self Inductance or The property of a coil due to which it opposes any change of current through it is known as self inductance. Its unit is Henry

$$\text{We have self inductance, } L = \frac{N\Phi}{I} \text{ or } L = \frac{\mu_0 \mu_r A N^2}{l} \text{ or } e_L = L \frac{dI}{dt}$$

Mutual inductance: It is the ability of one coil to produce an emf in a nearby coil by induction when the current in the first coil changes. This action is reciprocal. Its unit is Henry.

$$\text{We have mutual inductance, } M = \frac{N_2 \Phi_1}{I_1} \text{ or } M = \frac{\mu_0 \mu_r A N_1 N_2}{l} \text{ or } e_M = M \frac{dI_1}{dt}$$

Energy stored in a Magnetic field: The energy stored in a magnetic field is given by the expression,

$$E = \frac{1}{2} L I^2 \text{ Joules ; we have, } L = \frac{N\Phi}{I} \therefore E = \frac{1}{2} N\Phi I \text{ Also, } NI = H \times l \text{ and } \Phi = B \times A$$

$$\therefore E = \frac{1}{2} (H \times l) (B \times A) = \frac{1}{2} BHV \text{ Joules where } V \text{ is the volume of magnetic material}$$

Coils connected in Series:

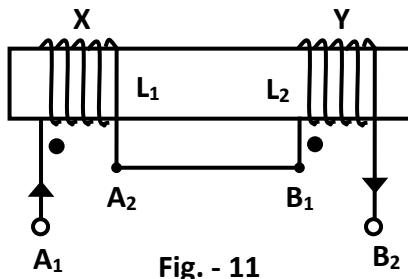


Fig. - 11

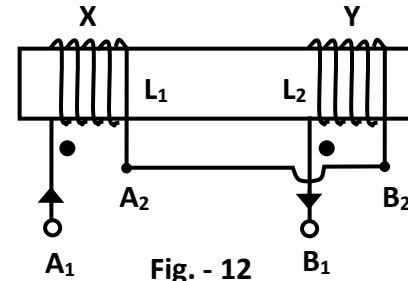


Fig. - 12

Let us consider two coils X and Y wound coaxially on an insulating cylinder with self inductances L_1 and L_2 Henry respectively and mutual inductance M Henry. Let the terminals A_2 and B_1 be joined together as shown in fig.-11. As per the dot convention as the currents are entering the dot, the fluxes produced in the coils are in the same direction and the coils are said to be cumulatively coupled. Hence, the equivalent self inductance L_A is given by $L_A = L_1 + L_2 + 2M$

Next, let the terminals A_2 and B_2 be joined together as shown in fig.-12. As per the dot convention, the fluxes produced in the coils are in opposition and the coils are said to be differentially coupled. Hence, the equivalent self inductance L_B is given by $L_B = L_1 + L_2 - 2M$

From the equations obtained for L_A and L_B , the value of Mutual inductance, $M = L_A - L_B / 4$

BASIC ELECTRICAL & ELECTRONICS ENGINEERING NOTES

Single phase transformer - Syllabus

General construction, working principle, emf equation, phasor diagram, equivalent circuits, voltage regulation, losses and efficiency, open circuit and short circuit test.

Transformer

A transformer is a stationary electrical apparatus that transfers power from one circuit to another by induction. It is the only A.C. machine that possesses the highest possible efficiency.

Construction: A transformer mainly consists of two windings and a laminated core with insulation. The winding connected to the supply is called primary winding and the winding connected to the load is called secondary winding. The winding made on the laminated core has a specific number of turns of copper conductors insulated from each other and from the core. The copper conductor can be of different cross sections. The core is made up of silicon steel laminations that are normally 0.3 to 0.6 mm thick. The laminations are insulated from each other by a thick coat of varnish. A bunch of laminations put together forms the core that are held together by bolts and nuts or riveted. High silicon content steel has a high relative permeability and low coefficient of Hysteresis. The use of high silicon content steel reduces hysteresis loss and by laminating the core the eddy current loss is minimized. The cores may be of different shapes. Depending on the shape of the core, we have core type transformers and shell type transformers. The figures show the magnetic cores used for shell type and core type transformers.

Core type: This type of core has a single window as shown in fig.-1 and the windings surround a considerable part of the core. The coils are of cylindrical type and are wound in helical layers. Each layer is insulated from the other by insulating paper. The windings are always interleaved to reduce leakage flux i.e. half the primary and half the secondary on one limb and the other halves on the other limb. The LV winding is always placed close to the core and the HV winding is made above LV winding. The individual laminations of the core may be a single piece or it may be made up of two pieces.

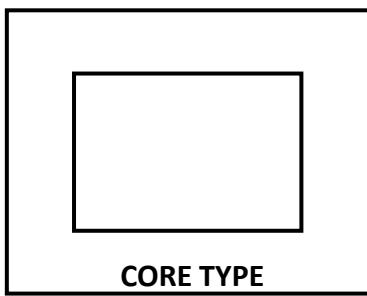


Fig. - 1

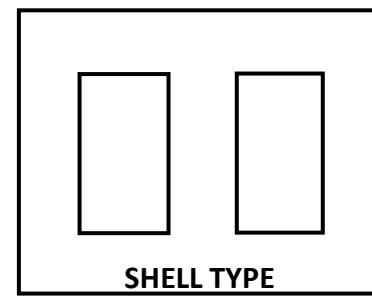


Fig. - 2

Shell type: This type of core has two windows as shown in fig. -2 and the core surrounds a considerable part of the winding. The individual laminations are usually made up of two pieces. The windings are always placed only on the central limb. Usually the HV winding is placed at the center and the LV winding is distributed equally on either sides of the HV winding. This type of winding is called Sandwich winding.

Principle of operation: The basic principle of operation of transformers is based on Faraday's laws of electromagnetic induction. The transformer has two windings wound on a laminated steel core. When AC supply is given to the primary winding, current flows through the primary winding producing a magnetic

BASIC ELECTRICAL & ELECTRONICS ENGINEERING NOTES

flux that is alternating in nature. As the magnetic field lines always form a closed loop the flux produced exists in the steel core. The flux that exists in the core links the primary as well as secondary winding. Therefore according to Faraday's law both the windings get emf induced in them.

Emf equation: Let an AC voltage V_1 of supply frequency f be applied to the primary winding.

The sinusoidal flux produced by the primary will be $\Phi = \Phi_m \sin \omega t$

The instantaneous value of emf induced in primary, $e_1 = -N_1 \frac{d\Phi}{dt} = -N_1 \frac{d}{dt} (\Phi_m \sin \omega t)$

$$e_1 = -\omega N_1 \Phi_m \cos \omega t = -2\pi f N_1 \Phi_m \cos \omega t \quad \therefore e_1 = 2\pi f N_1 \Phi_m \sin(\omega t - 90^\circ)$$

From the above equation, the maximum value of induced emf in primary is $E_{m1} = 2\pi f N_1 \Phi_m$

The rms value of the primary emf $= E_1 = 0.707 E_{m1} \quad \therefore E_1 = 0.707 \times 2\pi f N_1 \Phi_m = 4.44 f N_1 \Phi_m$

$E_1 = 4.44 f N_1 \Phi_m$ and $E_2 = 4.44 f N_2 \Phi_m$

Transformation ratios: Considering an ideal transformer, we have $V_1 = V_2$ and $E_1 = E_2$

The ratio of secondary voltage to primary voltage is called voltage transformation ratio.

It is represented by K and is also called as the turns ratio $V_2 = KV_1$

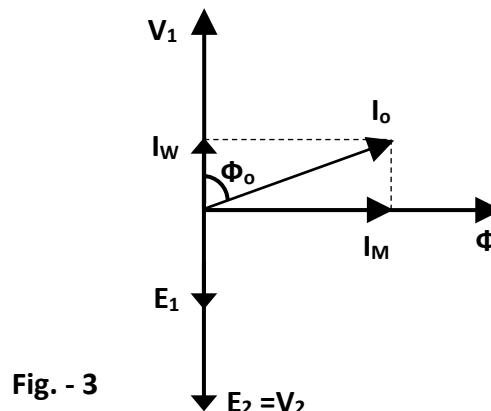
The voltage transformation ratio, $K = \frac{V_2}{V_1} = \frac{E_2}{E_1} = \frac{N_2}{N_1}$

In an ideal transformer, we also have, Output = Input or $V_2 I_2 = V_1 I_1$

$$\therefore \frac{I_2}{I_1} = \frac{V_1}{V_2} = \frac{E_1}{E_2} = \frac{N_1}{N_2} = \frac{1}{K} \quad \text{or} \quad I_2 = \frac{I_1}{K}$$

Hence, the currents are in the inverse ratio of the voltage transformation ratio.

Transformer on no-load: If the primary winding is applied with an AC supply voltage V_1 and the secondary winding is kept open, then the transformer is considered to be on no-load. Due to the applied voltage V_1 , an alternating current I_o flows in the primary that creates an alternating flux ϕ .



The applied voltage to the primary leads the flux by 90° . This flux that exists in the magnetic core links with the primary and secondary windings to induce emf's E_1 and E_2 that lag the flux by 90° . As there is no voltage drop in the secondary winding, $E_2 = V_2$. The no-load current I_o , also called as exciting current of the transformer lags the applied voltage V_1 by an angle ϕ_o . It is observed in fig. - 3, that the no-load current I_o has two components - I_M is the magnetising component that produces the desired flux in the core and is in

BASIC ELECTRICAL & ELECTRONICS ENGINEERING NOTES

phase with the flux ϕ . The other component I_w is called the working component or iron loss component that overcomes the hysteresis and eddy current losses occurring in the core. This component of the current is in phase with the applied voltage.

We have the magnetising component of the no-load current, $I_M = I_0 \sin \phi_0$

and the working component of the no-load current, $I_W = I_0 \cos \phi_0$

The no-load current, $I_0 = \sqrt{I_M^2 + I_W^2}$; The input power on no-load, $W_0 = V_1 I_0 \cos \phi_0$

The no-load power factor, $\cos \phi_0 = \frac{W_0}{V_1 I_0}$

Transformer on load : Whenever some variety of load is connected across the secondary winding of the transformer, the transformer is said to be on load. The load may be resistive, R-L or R-C. As the primary and secondary windings are made on the magnetic core using copper wires, let us assume that they possess a resistance of R_1 and R_2 respectively. The leakage flux in the primary winding varies linearly with the primary current and that in the secondary winding varies linearly with the secondary current. These leakage fluxes may be simulated by assigning primary and secondary leakage inductances along with the respective winding resistances. Let the reactance corresponding to the leakage inductance of primary and secondary be represented by X_1 and X_2 . Fig. - 4, represents the circuit of a transformer with its winding resistance and leakage reactance.

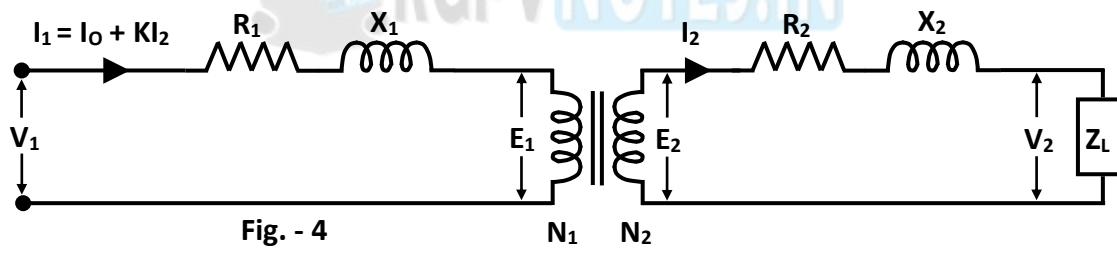


Fig. - 4

The value of R_1 and X_1 cause a voltage drop so that E_1 will be less than V_1 . Similarly, V_2 is less than E_2 due to R_2 and X_2 . Considering an R-L load, the current I_2 lags V_2 by an angle ϕ_2 . The primary current I_1 has two components, the no-load current I_0 and the current I_1^1 that neutralizes the demagnetizing effect of the secondary current I_2 . The additional mmf's due to the load currents in the secondary and primary windings create fluxes which cancel each other, leaving the original flux ϕ .

The magnitude of I_1^1 will be such that $N_1 I_1^1 = N_2 I_2$

$$\text{or } I_1^1 = \frac{N_2}{N_1} I_2 = K I_2$$

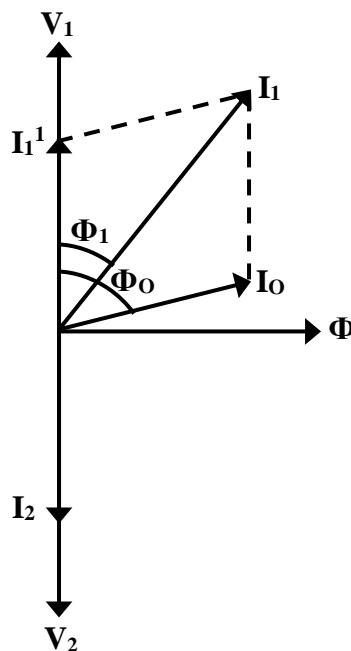
The primary current on load, $I_1 = I_0 + I_1^1$

$$\therefore V_1 = -E_1 + I_1 Z_1$$

$$\text{and } V_2 = E_2 - I_2 Z_2$$

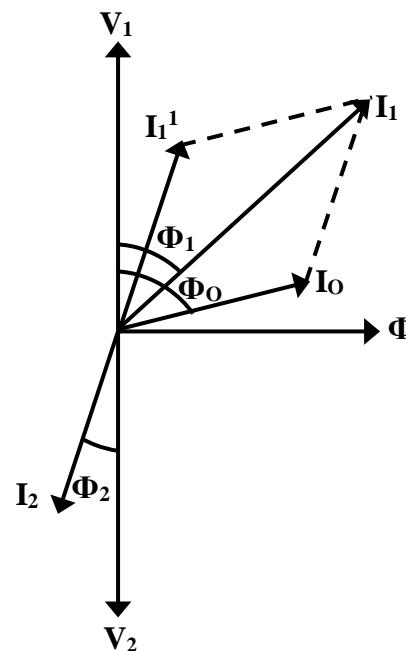
BASIC ELECTRICAL & ELECTRONICS ENGINEERING NOTES

The phasor diagram for the working of a transformer with different varieties of loads are shown below -



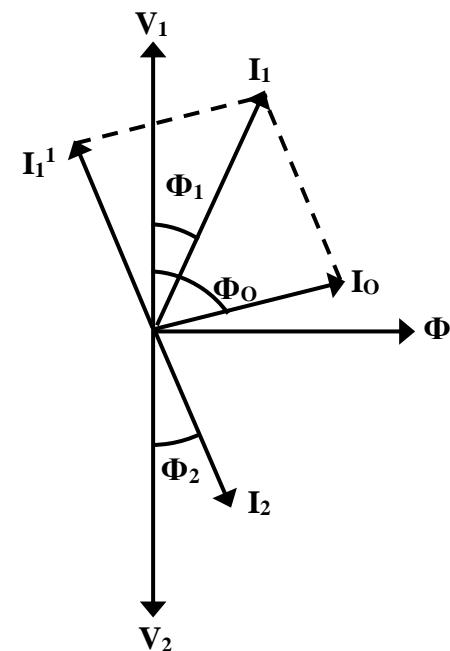
PHASOR DIAGRAM FOR TRANSFORMER WITH RESISTIVE LOAD

Fig. - 5



PHASOR DIAGRAM FOR TRANSFORMER WITH R-L LOAD

Fig. - 6



PHASOR DIAGRAM FOR TRANSFORMER WITH R-C LOAD

Fig. - 7

The applied voltage V_1 leads the magnetic flux ϕ by 90° and the secondary voltage V_2 is equal and opposite to V_1 . Let a R-L load be considered which draws a secondary current I_2 lagging V_2 by an angle ϕ_2 . The current I_1^1 represents the primary current that neutralizes the demagnetizing effect of secondary current I_2 . But $I_1^1 = K I_2$ is opposite to I_2 . I_0 is the no-load current of the transformer which lags the applied voltage by an angle ϕ_0 . The primary current I_1 is the phasor sum of I_0 and I_1^1 which lags the applied voltage V_1 by an angle ϕ_1 . The load power factor = $\cos \phi_2$, The primary power factor = $\cos \phi_1$

The input power to the transformer, $P_1 = V_1 I_1 \cos \phi_1$

The output power of the transformer, $P_2 = V_2 I_2 \cos \phi_2$

The phasor diagram for a transformer with resistive load is shown in fig.- 5, and phasor diagram for a transformer with R-L load is shown in fig.- 6. The phasor diagram for a transformer with R-C load is shown in fig.- 7.

Equivalent circuit of transformer:

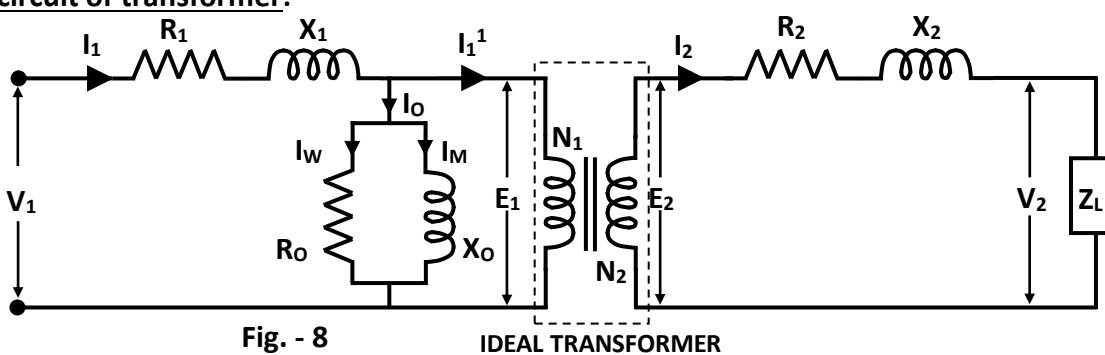


Fig. - 8

IDEAL TRANSFORMER

BASIC ELECTRICAL & ELECTRONICS ENGINEERING NOTES

An actual transformer has two electric circuits linked by a magnetic circuit. To simplify calculations, a transformer is often represented by its equivalent circuit as shown in fig. - 8. The effects of the core and the windings are represented by equivalent basic circuit elements and the transformer gets reduced to a simple circuit. An equivalent circuit is merely a circuit interpretation of the equations that describe the behavior of the device. The transformer windings are shown as ideal. The resistance and leakage reactance of the primary and secondary are shown separately in the primary and secondary circuits. The effect of magnetising current is represented by a reactance X_0 connected in parallel across the winding. The effect of core loss is represented by a non-inductive resistance R_0 . The no-load current I_0 in a transformer is only 1 to 3 % of its rated primary current; hence, it may be neglected, as it is not going to cause any serious error.

Approximate equivalent circuit of transformer: If the no-load current is neglected, we get the approximate equivalent circuit of the transformer. The equivalent circuit can be simplified by transferring the secondary resistance and reactance to the primary side in such a way that the ratio of E_2 to E_1 is not affected in magnitude or phase. If all the secondary quantities are referred to the primary, resistance and reactance are divided by K^2 , voltages are divided by K and currents are multiplied by K .

We get the equivalent circuit of the transformer referred to primary as shown in fig. - 9.

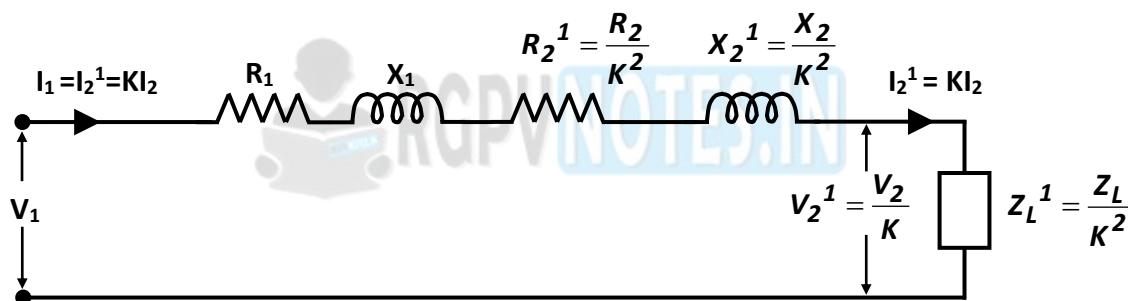


Fig. - 9

If all the primary quantities are referred to secondary, we get the equivalent circuit of the transformer referred to secondary. When primary quantities are referred to secondary, resistance and reactance are multiplied by K^2 , voltages are multiplied by K and currents are divided by K .

Voltage regulation of a transformer: It is defined as the change in the output terminal voltage of the transformer from no-load to load condition, expressed as a fraction or percentage of the no-load terminal voltage.

$$\% \text{ Voltage Regulation} = \frac{oV_2 - V_2}{oV_2} \times 100$$

Where oV_2 is the no-load secondary voltage and V_2 is the secondary voltage on load.

Losses in transformer: The losses occurring in a transformer are of two types-

1. Iron loss
2. Copper loss.

These losses appear in the form of heat and produce an increase of temperature and drop in efficiency.

BASIC ELECTRICAL & ELECTRONICS ENGINEERING NOTES

Core or Iron Losses: It comprises of Hysteresis and Eddy current losses and occur in the transformer core

due to the alternating flux. We have Hysteresis loss = $P_h = K_h B_m^{1.6} f V$ watts,

and Eddy current loss = $P_e = K_e B_m^2 t^2 f^2 V$ watts

Where K_h and K_e are constants, 'f' is the frequency of the supply, ' B_m ' is the Maximum flux density in the core, 'V' is the volume of magnetic material and 't' is the thickness of the lamination.

Both the losses depend on frequency and maximum flux density in the core. Since transformers are connected to constant frequency and constant supply voltage, both 'f' and ' B_m ' are constant. Hence, core or iron losses are practically the same at all loads.

Iron or core losses, P_i = Hysteresis loss + Eddy current loss = Constant losses.

Copper losses: These losses occur both in primary and secondary windings due to their ohmic resistance.

Copper loss = $P_c = I_1^2 R_1 + I_2^2 R_2$ = Variable losses

Where I_1 = Primary current, R_1 = Resistance of primary winding, I_2 = Secondary current,

R_2 = Resistance of secondary winding.

Total loss in Transformer = Iron loss + Copper loss = Constant loss + Variable loss = $P_i + P_c$

Efficiency of Transformer:

The efficiency of a transformer is defined as the ratio of output power to the input power.

$$\therefore \text{Efficiency} = \frac{\text{Output power}}{\text{Input power}} = \frac{\text{Output}}{\text{Output} + \text{Losses}} \quad \text{or} \quad \text{Efficiency } \eta_x = \frac{x.KVA.\cos\phi}{x.KVA.\cos\phi + P_i + x^2 P_c}$$

Where KVA is the output power rating of the transformer and 'x' is load factor,

$x = 1$ for Full load, $x = 0.5$ for Half full load and $x = 0.25$ for Quarter full load.

As Iron losses are independent of load they are considered as constant.

As Copper losses are proportional to the square of the load current, the copper losses = $x^2 P_c$

$$\% \eta_x = \frac{x.KVA.\cos\phi}{(x.KVA.\cos\phi) + P_i + x^2 P_c} \times 100$$

Transformer tests: The efficiency and voltage regulation of a transformer can be determined by two simple tests - Open circuit test and Short circuit test. These tests provide the required information using which the performance of the transformer can be ascertained without actually loading the transformer.

Open circuit test or no-load test: In this test the secondary winding is kept open and the rated voltage of the transformer is applied to its primary winding. The voltmeter measures the applied voltage V_1 . The ammeter measures the no-load current I_0 . The wattmeter measures the no-load input power W_0 . As the no-load current is very small and flows only in the primary, the copper losses due to it are negligible. Hence, the wattmeter reading practically gives the iron losses in the transformer. The circuit connection to perform this test is shown in fig. - 10.

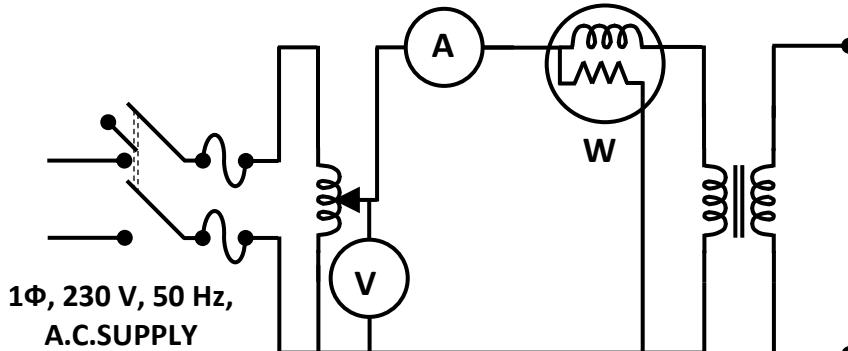


Fig. - 10

The input power on no-load = Iron losses = $P_i = W_o = V_1 I_o \cos\phi_o$

$$\text{No load powerfactor, } \cos\phi_o = \frac{W_o}{V_1 I_o}$$

Magnetising component of the no-load current, $I_M = I_o \sin\phi_o$

Working component of the no-load current, $I_W = I_o \cos\phi_o$

$$\text{Core loss component resistance, } R_o = \frac{V_1}{I_W} ; \quad \text{Magnetising Reactance, } X_o = \frac{V_1}{I_M}$$

Short circuit or Impedance test: In this test the secondary winding is short circuited and the voltage across the primary winding is adjusted such that the rated current flows through the primary winding. The voltmeter measures the applied voltage V_{1SC} . The ammeter measures the full load primary current I_1 . As the applied voltage is very small the iron losses in the core will be negligible. Under short circuit condition there is no output from the transformer. Hence, the wattmeter measures the full load copper losses W_s in the transformer windings. The circuit connection to perform this test is shown in fig. - 11

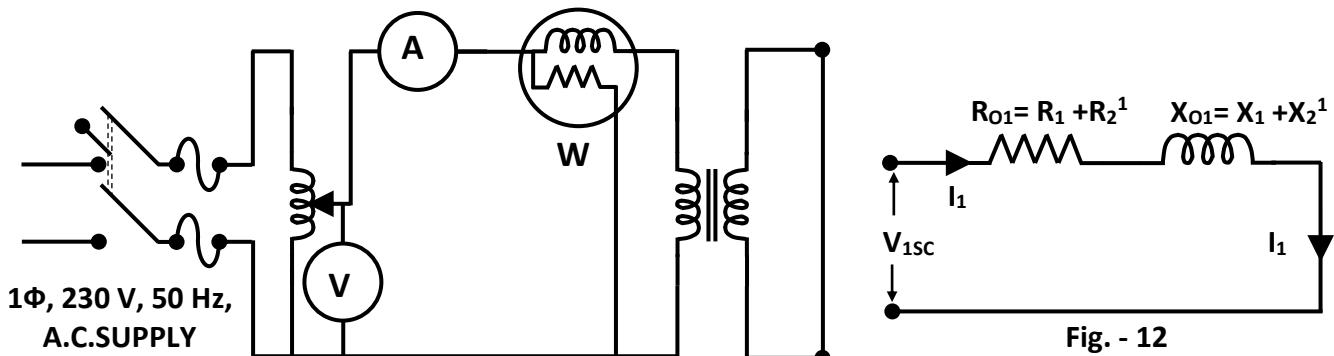


Fig. - 11

The equivalent circuit of a transformer on short circuit as referred to the primary is shown in fig. -12, as the no-load current is very small it is neglected.

The input power under short circuit condition = Copper losses = $P_C = W_s$

If total resistance of transformer referred to primary is R_{o1} ; total reactance referred to primary is X_{o1}

Then total impedance referred to primary is Z_{o1}

$$\therefore R_{o1} = \frac{W_s}{I_1^2} ; \quad Z_{o1} = \frac{V_{1SC}}{I_1} ; \quad X_{o1} = \sqrt{Z_{o1}^2 - R_{o1}^2} \quad \text{Short circuit power factor, } \cos\phi_2 = \frac{W_s}{V_{1SC} I_1}$$

BASIC ELECTRICAL & ELECTRONICS ENGINEERING NOTES

Relations for Efficiency and Voltage regulation: The efficiency of the transformer at any load factor 'x' and any power factor can be determined if the KVA rating of the transformer, Iron losses and the Full load copper losses are known. The relation given below helps in finding the efficiency from the OC and SC tests conducted on the transformer.

$$\% \eta_x = \frac{x.KVA.\cos\phi}{(x.KVA.\cos\phi) + P_i + x^2 P_c} \times 100$$

The voltage regulation of the transformer can be calculated by using the relation given below -

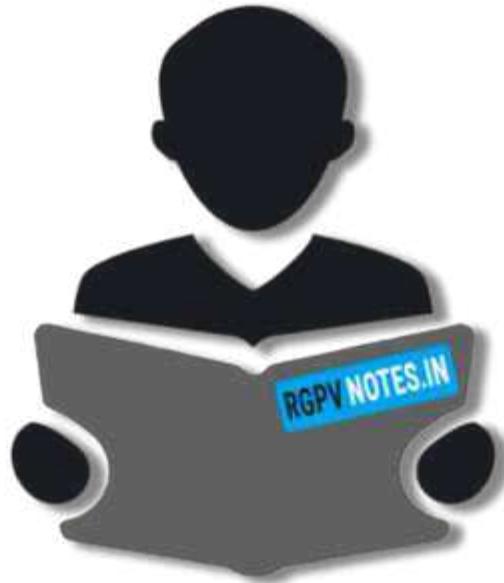
If the resistance and reactance of the transformer referred to the secondary side are known then

$$\% \text{ Voltage Regulation} = \frac{I_2 (R_{O2} \cos\phi_2 \pm X_{O2} \sin\phi_2)}{V_2} \times 100$$

(+ ve for lagging power factor and – ve for leading power factor)

V_2 is the secondary voltage of the transformer on no-load.





RGPV**NOTES.IN**

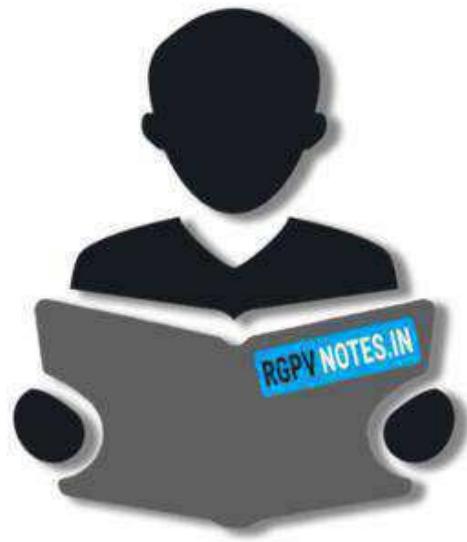
We hope you find these notes useful.

You can get previous year question papers at
<https://qp.rgpvnotes.in> .

If you have any queries or you want to submit your study notes please write us at
rgpvnotes.in@gmail.com



LIKE & FOLLOW US ON FACEBOOK
facebook.com/rgpvnotes.in



RGPVNOTES.IN

Program : **B.Tech**

Subject Name: **Engineering Chemistry**

Subject Code: **BT-101**

Semester: **1st**



LIKE & FOLLOW US ON FACEBOOK
facebook.com/rgpvnotes.in

ENGINEERING CHEMISTRY (BT-101) NOTES**UNIT III: LUBRICANTS & LUBRICATION****INTRODUCTION:**

All material surfaces, harder or softer, have many irregularities in the form of asperities (Peaks) and valleys. When two solid surfaces are pressed over each other, the asperities of both the upper surface and lower surface come in contact with each other, it causes some deformation effect on the metal surface which is more in case of ductile metals. This contact causes the weld junction which carries all the loads between the two surfaces. Thus, the true area of contact is only a small fraction of the apparent contact area between the two surfaces. Friction is the relative movement of two surfaces in contact with each other. Generally, there are two types of friction encountered, (i) sliding friction and (ii) rolling friction.

Surface wear during sliding is due to shearing of asperities

(i) Sliding friction: The friction caused by the sliding two materials of different hardness over one another. Due to this sliding, the peaks of the softer metal gets broken easily than the peaks of the harder metals.

The effects of the sliding friction are:

- (a) Ploughing i.e., cutting out the softer material by the asperities of the harder materials.
- (b) The interlocking of the surface irregularities.

(ii) Rolling friction: It occurs when a load sphere or cylinder rolls over a flat surface of the other body. The coefficient of the rolling friction is usually very low as compared to the sliding friction. Thus, the effect of rolling friction is much lower than the sliding friction, as the effect of rolling friction is caused by elastic deformation of the two surfaces. In case of sliding friction, the coefficient of friction is much larger in static condition than for the kinetic condition. Therefore, the lubricant plays less important role in case of rolling friction than the sliding friction.

In case of any friction a considerable amount of frictional heat is released at the rubbing surfaces which is more in case of sliding friction as the contact area of this friction is more. The frictional heat is not uniformly distributed over apparent contact area between the rubbing materials but it is highly localized, particularly at the surface asperities. If this frictional heat reaches the melting temperature of the material, welded junction may be formed. The amount of frictional heat depends upon the nature of the sliding material and the speed of the sliding.

Lubricant- A substance which is introduced between two relatively moving surfaces for reducing friction and wear is called lubricants. The process of reducing frictional resistance between moving surfaces by the application of lubricants is called lubrication.

Functions of Lubricants:

1. It reduces the loss of energy in the form of heat.
2. It increases the efficiency of machine.
3. It efficiently prevents the interlocking or inter-joints welding at the surface asperities.
4. It increases the smooth motion of the moving parts.
5. It reduces the surface deformation, wear and tear.
6. It reduces the expansion of metal by local frictional heat.
7. It protects the materials from corrosion.
8. Sometimes it stops the leakage of gases or pressures from the cylinder.

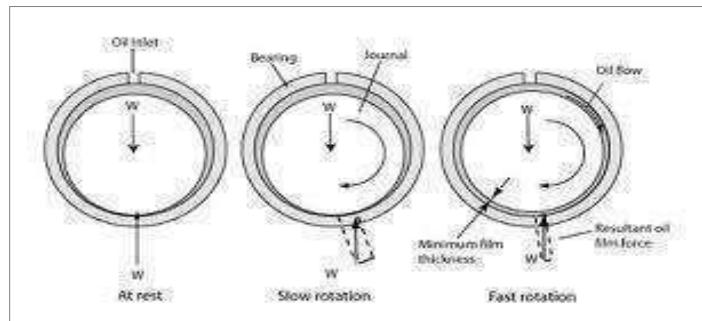
Mechanisms of Lubrication:-

There are mainly three types of mechanism of lubrication. They are:

1. **Fluid film or hydrodynamic lubrication.**
2. **Thin film or boundary lubrication.**

3. Extreme pressure lubrication.

(1) Fluid film or hydrodynamic lubrication: In this lubrication, two moving surfaces are separated from each other by a thick film of lubricating oil δ (thickness around 1000A^0). Hence it prevents the direct contact between these rubbing surfaces and welding of junctions generally does not occur. This happens because, the lubricating oil covers the irregularities of the Sliding surfaces & forms a thick layer in between them, so that there is no direct contact between the material surfaces. Actually the shaft floats in the lubricant.



The resistance to movement of moving parts is only due to the internal resistance between the particles of the lubricant moving over each other. Based on this principle, the lubricant chosen should have low viscosity. Selection of lubricant: In order to maintain a suitable viscosity of the oil, ordinary hydrocarbon lubricants are blended with long chain polymers. At operating condition, hydrocarbon petroleum fractions. Generally contain small quantities of unsaturated hydrocarbons, which get oxidized to form gummy products. To reduce the antioxidants like amino phenols are added. This type of hydrodynamic lubrication gives satisfactory results as the coefficient of friction in such cases is very low. Hydrodynamic friction occurs in the case of a shaft running at a fair speed with a low load.

(ii) Boundary lubrication or thin film lubrication: It is done when a thick fluid film cannot be maintained between the two moving surfaces so that direct metal to metal contact is possible. This happens when (i) the viscosity of the oil is low (ii) the load is very high (iii) the speed of the movement is very low and (iv) a shaft starts moving from rest. Under such conditions, the lubricant should be adsorbed physically or chemically on both the metallic surfaces. These adsorbed lubricant layers avoid the direct metal to metal contact. The coefficient of friction in these cases is slightly more than hydrodynamic lubrication. In boundary lubrication, surface asperities contact each other even though the lubricant supports much of the load. Friction depends mainly on the shearing forces necessary to cleave these adhering asperities and wear and friction can be reduced by certain additives. Wear inhibitors and lubricity agents are polar materials that adsorb on a metal and provide a film that reduces metal-to-metal contact.

Condition of lubricants: The lubricant used for boundary lubrication should have:

- (1) Long hydrocarbon chains.
- (2) Lateral attraction between the chains.
- (3) Polar groups to promote spreading and orientation over the metallic surfaces at high pressure.
- (4) Active groups to form chemical bonds with metallic surfaces.

In this boundary lubrication the metal surfaces are closer to each other, but still they are well separated by the lubricant. The property of the oil which enhances the adsorption and maintenance of the thin film is called oiliness. High viscosity index, resistance to heat and oxidation, good oiliness and low pour

ENGINEERING CHEMISTRY (BT-101) NOTES

point are some of the good qualities of boundary lubricants. Boundary lubrication (i.e., the boundary friction) boundary lubrication is the lubricant from the friction surfaces between the molecules and the internal friction between molecules (i.e., liquid lubrication) the transition to direct contact with the surface friction before the critical state. At this time there is a layer of friction interface adsorption film thickness is typically about $0.1\mu\text{m}$, with some lubrication. We call this layer of thin boundary film. Boundary lubrication film depends on the nature of the friction surface; depends on the oil lubricant additives, extreme pressure additives on the friction surface of the metal structure of the boundary film formation with the oil viscosity has little mouth. Generally vegetable and animals oils have all the above mentioned good qualities of lubricant. They are adsorbed (physically or chemically) easily on the, metal surfaces with the active group of -COOH. But they tend to break down at high temperatures. In order to improve the oiliness of the mineral oils, usually little fatty acids are added.

(iii) Extreme pressure lubrications: Extreme pressure lubrication are a special case of boundary lubrication, friction is Deputy Commissioner in the override (or high contact stress), speed, temperature conditions, extreme pressure lubricant additives react with the metal friction surface generate a chemical reaction film, separated the two friction surfaces, and play a lower coefficient of friction, reduce wear and tear (or change the direct contact with the metal surface of serious wear and tear), to the role of lubrication, it is called extreme pressure lubrication. When the moving surfaces are under very high pressure and speed, the lubricant may decompose or vaporize or they may not stick on the surfaces because high local temperature is produced. To meet these extreme pressure conditions, special additives called "extreme pressure additives" are added to the mineral oils. Extreme pressure additives are compounds having active groups such as chlorine (chlorinated esters), sulphur (sulphurised oils) and phosphorus (tricresyl, phosphate). At high temperatures, they react with metal giving surface layers like metallic chlorides, sulphides or phosphides. These surface layers have high melting point and serve as good lubricant under extreme pressure and temperature conditions. Extreme pressure (EP) additives are a special class of boundary lubrication additive which react with the metal surface to form compounds with lower shear strength than the metal.

Classification of Lubricants:-

Lubricants may be broadly classified, on the basis of their physical state as:

1. Liquid Lubricants.
 - (a) Vegetable oils and animals oils.
 - (b) Mineral oils from petroleum.
 - (c) Blended oils or doped oils or compound oils.
 - (d) Synthetic oils.
2. Semi solid lubricants.
3. Solid lubricants.

Lubricating Oils/ Liquid lubricants:

(a) Vegetable or animal oils: These are most commonly used lubricants. They possess good oiliness and adsorb themselves on the metallic surfaces. However, they decompose at high temperatures and undergo oxidation easy, forming gummy and -acidic hydrolyzed products and get thickened on coming in contact with air. Thereby restricting the smooth movement of the moving surfaces. To overcome such restrictions, usually they are blended with mineral oils (blending agents).

(b) Mineral oils: They are obtained by fractional distillation of petroleum. They are the heavier fractions containing long chain hydrocarbons in petroleum oils ranging between 12 to 50 carbon atoms. The shorter chain oils have lower viscosity than the longer chain hydrocarbons. They are widely used as

ENGINEERING CHEMISTRY (BT-101) NOTES

lubricants as they possess good stability under service conditions but poor oiliness. The oiliness of these oils can be improved by adding oils like oleic, stearic acids etc. They have impurities like sulphur compounds, gum forming alkenes, asphalt, waxy impurities, other color impurities, reluctant and some other unwanted impurities. All these impurities should be removed to prevent the formation of corrosive and gummy products to decrease the friction.

Removal of impurities: A number of processes are used for removing these undesired impurities as follows:

(i) Solvent refining: Sulphur compounds, alkenes and other gum products can be removed by this process. In this process, the oil is mixed with a suitable solvent like phenol which is immiscible with oils. But the undesirable impurities are highly soluble in phenol, thereby the liquid separates into two layers (oil goes up and phenol comes down).

Oil layer is free from impurities but containing some solvents and the solvent layer contains the impurities. Then the oil layer is distilled out and some additives may be added to improve the lubricating quality of the oil.

(ii) Dewaxing: Waxy impurities are removed by treating the lubricants with suitable solvents like propane and then the wax solution is cooled. The wax precipitates, which is removed from the oil by passing the oil wax suspension through a filter. The solvent present in the oil is then recovered by distillation.

(iii) Filtration: The colored impurities can be removed by filtration through bauxite or clay.

(iv) Acid refining: Sometimes the dewaxed oils contain a number of undesirable impurities. To eliminate these impurities, the dewaxed-oil is treated with concentrated H_2SO_4 and then agitated so that the impurities dissolve in acid and it is filtered out. This filtrate is neutralized by adding a calculated amount of a base NaOH.

(c) Blended lubricating oils: Single oil cannot possess all the good qualities of the lubrication. Therefore to get the satisfactory lubricant with desirable characteristics, specific additives are blended with the oil. These are called blended oils. An additive is a material that imparts a new or desired properties to the lubricating oil. It may also enhance a desirable property that the lubricating oil already possess to some degree. Broadly, there are two types of additives chemically active additive & chemically inert additives. Chemically active additive are those which chemically interact with metals (to form protective films) .dispersant, detergents, extreme pressure agents, etc. Chemically inert additives are those additives which improve the physical properties that are critical to the effective performance of the lubricant. Like viscosity index improver, foam inhibitors etc.

Greases or Semi solid lubricants:-

The most important semisolid lubricants are Greases and Vaseline.

Grease: Grease is used to lubricate journal bearings when cooling of the bearing is not a factor, typically if the bearing operates at relatively low speeds. Grease is also beneficial if shock loading occurs or if the bearing frequently starts and stops or reverses direction. Grease is almost always used to lubricate pins and bushings because it provides a thicker lubricant than oil to support static loads and to protect against vibration and shock-loading that are common in many of these applications. Lithium soap or lithium complex thickeners are the most common thickeners used in greases and are excellent for most journal bearing applications.

Grease is a semisolid combination of a petroleum product and soap. It is obtained by saponification of fat (fatty acids etc.) with alkali like NaOH followed by adding hot Lubricating oil under agitation. The amount of mineral oil added determines the consistency of the finished grease. Soaps are gelling agents which can interconnect with the added oil. Therefore, the structure of the lubricating greases is like gel.

ENGINEERING CHEMISTRY (BT-101) NOTES

Soaps dissolve in the oil at higher temperatures so that the interconnecting bonds will be broken out and the grease liquefies. To improve the heat resistance of grease; inorganic solid thickening agents like clay, colloidal silica, carbon black etc. are added to it.

Advantages of greases:

1. Greases have higher frictional resistance than oil.
2. Because of the presence of soap in oil, greases stick well on the surfaces.
3. Greases are used in situations where oil film cannot be remaining in its position i.e. in machine working under high pressures at slow speed.
4. Greases can support heavier load at low speed.
5. Greases can be used in bearing and gears that work at high temperatures.
6. They are used in situations where sealing is necessary against entry of dust, dirt, grit or moisture.
7. Greases are used in situations, where dripping or spouting of oil is undesirable; therefore it can be used in textile and good products manufacturing industries.
8. They also do not require much attention as oils.

Disadvantages of greases:

1. They have high coefficient of friction.
2. They have a tendency to separate into oils and soaps. .
3. On long use, oil in the grease may evaporate.
4. Greases cannot effectively dissipate heat from the bearing.

Greases are classified based on the soap used in their manufacture as below:

- (a) **Lime or calcium soap grease or cup greases:** They are the emulsions of petroleum oil with calcium soaps. They are insoluble in water, so water resistant. Above 65°C they cannot be used because oil and soap separates out above this temperature. They are widely used as they are the cheapest.
- (b) **Sodium soap greases:** They are sodium soaps dispersed in petroleum oils. They are water soluble, so they are not water resistant. They can be used up to 175° C because of its high melting point and fibrous structure. They are suitable for use where the lubricants get heated due to friction.
- (c) **Lithium soap greases:** They are petroleum oils, thickened by mixing lithium soaps. They are water resistant, high temperature withstanding, good mechanical stability and low oxidation tendency. They are suitable for use at low temperature about 15°C only. They are used in multipurpose.
- (d) **Aluminum soap greases:** They are aluminum soaps dispersed in oils. Due to its low soap content, they are water proof. They cannot be used above 90°C.
- (e) **Axle greases:** They are the resin greases. They are separated by adding slaked lime to the mixture of soluble resin oil and fatty oils. They are allowed to react at 58°C. The mixture is thoroughly mixed and allowed to stand, when grease flats as stiff mass called cold set grease. Filters like talc and mica are also added to these greases. They are water resistant and suitable for use as axle greases and for heavy equipments.

Solid Lubricants: Solid lubrication between the friction surface materials into the solid lubricant powder also can play a good lubricating effect. Friction between the two surfaces of solid lubricant, its shear resistance is very small, slight external force, will have a slip between molecules. This put the two outside the friction between the grinding into a solid lubricant in friction between molecules. There are two necessary conditions for solid lubricant, the first is a solid lubricant molecules should have a low shear strength, it is easy to slip; followed with solid lubricant to the friction surface has a strong affinity, in the friction process, always has maintained that the surface friction layer of solid lubricant, and a layer of solid lubricant which does not corrode the surface friction. Generally attached to the metal surface, a mechanical, but there are also form a chemical combination. The nature of the solid material with many, such as graphite, molybdenum disulfide, talc and so on. For non-layered structure for the

ENGINEERING CHEMISTRY (BT-101) NOTES

solid lubricant, or soft metal, mainly for its low shear stress, play a role of lubrication, then it attached to the friction surface lubricating film. For the already formed solid lubricating film lubrication mechanism of boundary lubrication mechanism can explain the lubrication approximation.

The two most usual solid lubricants employed are graphite and molybdenum disulphide. These lubricants are used either in the dry powder form or mixed with water or oil. The solids fill up the low spots in the surfaces of moving parts and form solid films, which have low frictional resistance. The usual coefficient of friction between solid-lubricants is between 0.005 and 0.01.

Advantages of solid lubricants:

1. They are used in heavy machineries working at very high loads and slow speeds.
2. The operating temperature is too high.
3. Liquid lubricants and greases are easily contaminated with dust which is unaccountable, for example, in commutator blades of electric motors and generators, because proper films cannot be maintained.
4. Combustible lubricants must be avoided.

I. Graphite: Graphite is most widely used of all solid lubricants. In graphite, carbon atoms are arranged in hexagons in several flat layers, which are held together by only weak bonds so that the force to shear the crystals parallel to the layers is low.

Each carbon atom in a layer is surrounded by three other carbon atoms.

Lubricating action of graphite: It is very soapy, non flammable graphite is mainly used at high temperature in the absence of air and high pressures. Graphite is used either in dry powder form or dispersed in oil (oil dag) or in water (aqua dag) or as graphite grease. Oil dag is found particularly useful in IC engines, because it forms a film between the piston rings and the cylinder and gives a tight fit contact, thereby increasing compression. Water dag is used in food industries, where a lubricant free from oil is needed. Graphite grease is used at high temperature.

Uses: It is used as lubricant in air compressors, lathes, general machine-shop works, food stuffs industry, railway track joints etc.

2. Molybdenum disulphide: It has sandwich like structure 'Mo' and 'S' are separately arranged in different layers and are sandwiched alternatively.

Lubricating action of molybdenum disulphide: Poor inter laminar attraction is responsible for low shear strength in a direction parallel to the layers. It is stable up to 400°C and is used at high temperatures. It possesses very low coefficient of friction. It can be used as a dry powder or oil-dag or aqua-dag or grease. A solid film made from 70% MoS₂, 7% graphite and 23% silicate is used in space vehicles which can withstand high temperature and even nuclear radiation. The other commonly used solid lubricants are talc, mica, soap stone etc. Teflon acts as a lubricant in gear pumps, periscope etc.

Synthetic Lubricants: These are synthesized specially to meet the severe operating conditions. The synthetic lubricants can perform well over a wide range of temperatures from -50°C upto 260°C. These lubricants possess low freezing points, high viscosity index and non flammable. Polyalkene glycols, silicones, chlorinated and fluorinated hydrocarbons; organic amines, imines and oxides are the important synthetic oils.

Advantages of synthetic lubricants

1. They possess high thermal stability at high operating temperatures.
2. They have viscosity index and high flash points.
3. They are chemically stable and having low freezing points.

ENGINEERING CHEMISTRY (BT-101) NOTES

(a) Polyalkene glycols: Polyalkene glycols like polyethylene glycol, poly-propylene glycol etc. can be used as both water soluble and water insoluble lubricants in rubber bearings and joints. They have all the three advantages mentioned above.

(b) Silicones: These synthetic lubricants are not oxidized below 200°C and possess high viscosity index. At temperature above 200°C , silicones are oxidized quickly and undergo cracking process at about 230°C . Therefore such lubricants are used for low temperature lubrication purposes but not employed for high temperature applications.

(c) Chlorinated and fluorinated hydrocarbons: They are not decomposed by heat, not easily oxidisable and chemically inert and resistant to chemicals etc.

(d) Organic amines: They are good synthetic lubricants, since they possess low pour points and high viscosity index. They can be used under temperature conditions of -50°C to 250°C .

Lubricating Emulsions: - An emulsion is a heterogeneous system consists of disperse phase and dispersion medium. In this emulsion, the two phases are immiscible liquids. The liquid being dispersed as fine droplets (dispersed phase) into a fairly coarse dispersion medium. The size of dispersed phase should be within the range of 1 micron to 6 micron. As the two immiscible liquids is inherently unstable, it is prepared by vigorous stirring and addition of emulsifier or emulsifying agent to make it more stable. The emulsifying agents contain polar and non-polar groups and they are classified as hydrophilic end and hydrophobic end depending on their affinity towards water molecule. The hydrophilic end emulsions have greater affinity towards water while the hydrophobic ends do not have any affinity towards water, rather they are preferred to wet by oil. The emulsifier molecule is adsorbed at the interface of the two phases, resulting in formation of a protective film around the dispersed droplet. For example, soaps of different types, detergents, long chain sulphonated acids, lyophilic colloids etc. act as emulsifying agents in different cases. A certain large number of droplets of a liquid into another liquid increases the surface area and hence requires large amount of energy. The required energy would be less if the surface tension is decreased. The function of the emulsifying agents is to diminish the surface tension.

Emulsions are broadly classified into two types (I) Oil in water type in which the disperse phase is the oil in water medium (2) Water in oil type in which water droplets are dispersed in oil. The type of emulsion produced depends upon the nature of the emulsifying agents used. For example, when sodium oleate is used as an emulsifier, oil in water emulsion is formed. But if soap with a bivalent cation is employed, say calcium oleate, water in oil type emulsion is obtained. In order to ascertain the type of given emulsion, a small quantity of water is added with stirring and the effect is observed under a microscope. If the emulsion is oil in water type, water would mix freely with the excess medium, the droplets remaining undisturbed. If it is water in oil type, the addition of oil to the emulsion would produce similar effect.

Properties and Testing of lubricating oils

1. Viscosity and Viscosity index
 2. Flash Point and Fire Point
 3. Cloud Point and Pour Point
 4. Aniline Point
 5. Steam Emulsion Number
 6. Neutralization Number
 7. Saponification Number
 8. Iodine value
 9. Carbon residue
-

1. VISCOSITY: - *It is the property of a fluid that determines its resistance to flow. It is an indicator of flowability of a lubricating oil, the lower the viscosity, greater the flowability. It is mainly, due to the forces of cohesion between the molecules of lubricating oil.*

Absolute viscosity may be defined as "the tangential force per unit area which is required to maintain a unit velocity gradient between two parallel layers; It is denoted by eta (η). Its unit in C.G.S. System is poise and its dimensions are $ML^{-1}T^{-1}$

Absolute Kinematic viscosity is the ratio of absolute viscosity to 'density for any fluid. It is denoted by v. Its unit in C.G.S. system is stokes and its dimensions are L^2T^{-1}

Effect of temperature on viscosity.

Like any other fluid, viscosity of lubricating oil is inversely proportional to temperature. *i.e.,* with increase of temperature, viscosity decreases. This is due to the decreases in intermolecular attraction. At higher temperatures oils must have sufficient viscosity to carry loads. Hence, heavier oils are used at higher temperatures. Similarly, light oils are used at low ambient temperatures.

Effect of pressure on Viscosity

Lubricating oils are subjected to extreme pressure at the interface between gears and between rolling element and race in a rolling bearing.. At such high pressures, viscosity of lubricating oils increases considerably.

Significance of Viscosity.

Viscosity helps in the selection of good Lubricating oil. For instance, Light oils have low densities and easy flowability. These oils *i.e.,* generally used on parts moving at high speed promotes the formation of a good oil film. Moreover, light oils do not impose much drag on high-speed parts.

Diagram (from manual):

In contrast, Heavy oils are used on parts moving at slow speed under heavy loads as they resist being squeeze out. Light oils are not suitable in this case as they have low viscosities and it is not possible to maintain lubricant film between the moving surfaces. Subsequently, excessive wear will occur. To sum up, in hydrodynamic lubrication, that lubricant is selected which should have a sufficient high viscosity to adhere to the bearing and resist being squeeze out due to high pressure and get fluid enough to resist excessive friction due to the shearing of oil itself. Thinner oil can easily dissipate the frictional heat because of good oil circulation.

Measurement of viscosity of lubricating oil

There are many instruments for measuring viscosity and are known as viscometers (or viscosity meters). For examples, Kinematic & Saybolt universal viscometers used in the United States and the Redwood viscometers are commonly used in England.

The Saybolt Viscometer

It consists of cylindrical brass cup in the bottom of which is an orifice of specified dimension. This cup is surrounded by constant temperature oil bath. A desired bath temperature can be obtained by adjusting the temperature regulator. When the sample of lubricating oil reaches test temperature. The time required for 60 mL of the oil to run through the orifice is measured. The oil sample flow by gravity under a standard falling head and at a temperature of 100°F and 210° F. A calibrated standard flask collects the liquid sample. The time of efflux is measured in seconds and are reported as Saybolt Universal Seconds (SUS); for example, 260 SUS, at 100° F. For heavy lubricating oils with high viscosities, a large orifice is used in the same apparatus, and the results are reported in Saybolt Furol Seconds (SFS) at a Specified temperature.

The kinematic Viscometer

It is a U-type of glass apparatus having respectively cone and two bulbs at its two sides. At the bottom portion of two bulb side, capillary tube is attached. It is used for the determination of Kinematic

ENGINEERING CHEMISTRY (BT-101) NOTES

viscosity. The four simple measurements steps are shown. For a fixed volume of sample, time is measured for the sample to flow through a calibrated capillary under an accurately reproducible head of liquid and at a constant temperature. From the measured efflux time, the kinematic viscosity is calculated.

$$\text{Kinematic viscosity in centistokes (CST)} = C \times t$$

Where C = Viscometer constant and.

t = Observed flow time in seconds.

The Redwood Viscometer: It is of two types:

(a) **Redwood viscosity No. I-Universal** and (b) **Redwood viscosity No.2-Admiralty.**

The essential differences between the two are:

	(a)	(b)
Dimensions of Orifice.	Length: 10 mm Diameter: 1.62 mm	50 mm 3.80 mm
Useful for	Low viscous oils	Higher viscous oils
Receiving flask has	Smaller mouth	Larger mouth

The Redwood No. 1 apparatus consists of a cylindrical brass oil cup (90mm in height and 46.5 mm in diameter) that holds the test sample of lubricating oil. Bottom of the oil cup is fitted with a polished-agate discharge tube containing an orifice of specified dimension. The oil cup is surrounded by water bath for adjusting the temperature. A calibrated receiving flask (known as Kohlrausch flask) is provided for receiving the oil from polished-agate discharge tube. It is shown in when the sample reaches test temperature the time for 50mL of the sample flow through the orifice is measured. Results are reported in seconds.

For example, Redwood Viscosity No. 1 at 140°F , 350 seconds.

Conversion Formulas:

Let time of efflux = t seconds,

Saybolt: $cSt = 0.22 t - 180ft$

Redwood: $cSt = 0.26 t - 171ft$.

Significance of viscosity measurements.

Viscosity is the property of lubricating oil that determines its ability to lubricate. Viscosity values are used in evaluating load carrying capacity, in denoting the effect of temperature changes, for establishing uniformity in shipments and for determining the presence of contaminants in used oil during absolute Viscosity values are required for use in all-bearing design calculations and other lubrication Engg. Technical design problems. Kinematic values are applied in oil blending procedures.

Viscosity Index -With changes in temperature, the viscosity of lubricating oil varies, the higher the temperature, the lower the viscosity and vice-versa. The rate of variation of Viscosity with temperature is different for different base oils or fluids. For example, in petroleum oils, the viscosities of naphthenic base oils vary more over the same temperature range than those of paraffinic base oils. The rate at which the viscosity of oil changes with temperature is measured by an empirical number, known as the **viscosity-index (V.I.)**. A relatively small change in viscosity with temperature is indicated by high viscosity index. Whereas a low viscosity index shows a relatively large change in viscosity with temperature. The Pennsylvanian oils, consisting mainly of paraffin's, are arbitrarily assigned a viscosity index value of 100 as they exhibit a relatively small change in viscosity with a rise in temperature. Oils of Gulf-coast origin, consists mainly of alicyclic (naphthenic), and are arbitrarily assigned a viscosity index value of 0 as they exhibit a larger change in viscosity with a rise in temperature. In industry, lubricating oils of high viscosity index are preferred since they have practically the same viscosity over a range of temperatures. Determination of Viscosity-index, Viscosity index of test oil is determined with

ENGINEERING CHEMISTRY (BT-101) NOTES

the help of two types of standard oils viz. Pennsylvanian oils and Gulf oils having V.I.'s 100 and 0 respectively. The V.I of the test oil is given by the formula:-

$$\text{V.I.} = \frac{L - U}{L - H} \times 100$$

V.I. = viscosity index of the oil under-test.

L = viscosity at 100°F of the low-viscosity standard oil having a V.I. of a (*i.e.*, Gulf oil) and also having the same viscosity as the oil under test at 210°F .

U = Viscosity at 100°F of the oil under test.

H = viscosity at 100°F of the high-viscosity standard oil having a Viscosity of 100 (*i.e.*, Pennsylvanian oil) and 50 having the same viscosity as the oil under test 210°F .

Example 1. An oil of unknown viscosity-index has a Saybolt universal viscosity of 60 seconds at 210°F and of 600 seconds at 100°F . The high viscosity index standard (*i.e* Pennsylvanian) oil has Saybolt viscosity of 60 seconds at 210°F and 500 seconds at 100°F . The low viscosity-index standard (*i.e.*, Gulf oil) has a Saybolt universal viscosity

6? Seconds at 210°F and 800 seconds at 100°F . Calculate the viscosity index of unknown oil.

Solution. Here $L = 800\text{ s}$, $H = 500\text{ s}$ and $U = 600\text{ s}$

So, viscosity-index of unknown oil $\text{V.I.} = \frac{L - U}{L - H} \times 100$

$$800 - 600 / 800 - 500 \times 100$$

$$\text{V.I.} = 66.67$$

Example 2. An oil sample under test has a Saybolt universal viscosity of 64 sec. at 210°F and 564 seconds at 100°F . The low viscosity standard (Gulf oil) possess Saybolt viscosity of 64 seconds at 210°F and 774 seconds at 100°F and 414 sec. at 100°F . Calculate the viscosity-index of the oil sample under test.; Solution. Here, $L = 774\text{s}$, $H = 414\text{s}$ and $U = 564\text{s}$

So, viscosity index of the oil-sample. Under test, $\text{V.I.} = \frac{L - U}{L - H} \times 100$

$$= 774 - 564 / 774 - 414 \times 100$$

$$\text{V.I.} = 774 - 414 \times 100 = 58.33$$



Viscosity Index and Molecular Structure of Oil

There is a direct co-relation between molecular structure of lubricating oil with its viscosity and viscosity-index. A high V.I. is exhibited, by those lubricating oils which have linear or rod-like shaped molecules of higher molecular weights. This is due to the greater inter-molecular attraction.

Viscosity-Temperature Curves.

The variation of viscosity with temperature can also be indicated by viscosity- temperature curves: In fact, viscosity index is the numerical expression of the average slope of the viscosity-temperature curve of lubricating oil between 100°F and 210°F . Lubricating oils with small variation in viscosity with temperature exhibit flatter viscosity-temperature plots and they have high VI's.

GRAF (from manual)

1. Flash and Fire & Point:

The *flash point* of oil is the lowest temperature at which it gives off vapors that will ignite for a moment when a small flame is brought near it. The *fire point* of oil is the lowest temperature at which the vapors of the oil burn continuously for at least 5 seconds when a small flame is brought near it.

The flash points and fire points are used to indicate the fire hazards of petroleum products arid evaporation losses under high temperature operations. Knowledge of flash and fire points in lubricating oil aids in precautionary measures against fire hazards. A good lubricant should have flash point at least above the temperature at which it is to be used.

Measurement of Flash and Fire Points of Lubricating Oil

(i) **The Pensky-Marten Closed cup method.** It is used to determine- the flash Point of lubricating oils, fuel oils, solvents, solvent containing materials and suspension of solids, except cut-back asphalt. It consists of a *cup* made of brass, which is about 5.5 mm deep and 5 cm in diameter. The lid of the cup is provided with four openings of standard sizes. Through one of these openings passes a *stirrer* carrying

ENGINEERING CHEMISTRY (BT-101) NOTES

two brass blades; while the second opening is meant for admission of air. Through third opening passes a thermometer. While the fourth is meant for introducing test flame. At the top of the cup, a *shutter* is provided. By moving the shutter, opening in the lid opens and flame (from *flame exposure device*) is dipped into this opening. There by bringing the flame over the oil surface. As the test-flame is introduced in the opening, it gets extinguished, but when the test-flame is returned to its original position, it is automatically lighted by the pilot burner. There is also a stove consisting of an air bath and a top plate on which the flange of the cup rests. The air bath may be either a flame-heated metal casting or an electric-resistance element. Pensky-Marten's flash point apparatus is shown in Fig. **DIGRAM (from manual)**

Procedure. The oil sample is filled up to the specified filling mark in the cup. It is then covered and positioned properly in the stove. The thermometer is inserted in the sample. The test flame is lighted, while being stirred, the sample is heated at the rate of 9° to 11° F per minute. Upto 220° F, the test flame is applied at every 2° F rise in temperature, thereafter it is applied at every 5° F rise in temperature. When a distinct flash occurs in the interior of the cup at the time of the flame application, the temperature reading on the thermometer is the flash point.

2. Cloud and Pour Points.

The *cloud point* of petroleum oil is the temperature at which solidifiable compounds, like paraffin wax, present in the oil begin to crystallize or separate from solution. When the oil is cooled under specified conditions. Naphthenic type of oils that are quite wax-free show no cloud points. The *pour point* of a petroleum oil is the temperature at which the oil ceases to flow or pour.

At the Cloud point, oil becomes cloudy or hazy in appearance. Cloud and pour-points indicate the suitability of lubricants in cold conditions. Lubricants used in a machine working at low temperature should possess low pour-point; otherwise solidification of lubricant will cause jamming of the machine. Pour point is of importance in establishing the lowest temperature at which a diesel fuel is still sufficiently fluid to be pumped or transferred. Oil with a low pour point should be selected whenever the oil must remain fluid at low temperatures.

Determination of cloud and Pour Points- These determinations are carried out with help of apparatus. The apparatus consists of a flat-bottomed tube (about 2 cm high and 3 cm in dia.) enclosed in an air-jacket. The air-jacket is surrounded by freezing mixture. (Ice + NaCl) contained in a jar.

To determine cloud point, a sample of the lubricating oil (moisture free) is poured into a test jar and cooled in progressive steps. When inspection first reveals a distinct cloudiness or haze at the bottom of the test jar, the temperature is recorded as the cloud point.

To determine pour point, a sample of oil is cooled in flat-bottomed tube (*i.e.*, test jar) under specified conditions; the temperature is observed in increments of 5° F until no movement is observed at the surface of the oil when the tube is held in a horizontal position for 5 seconds. This temperature is recorded as the solid point. By definition the pour point is 5° F above this temperature.

Significance. Cloud point is useful for estimating the temperature at which filter screens in the fuel intake system of diesel engines might become dogged because of wax separation. Filterability depends on type of wax, micro crystalline or amorphous. The amorphous wax is sticky in nature and would more easily clog filter screens. Pour point values of petroleum and non-petroleum lubricants are significant as many operations must function in sub-freezing conditions.

Aniline Point: Aniline point is defined as "the minimum equilibrium solution temperature for equal volume of aniline and oil sample." It is determined by thoroughly mixing equal volumes of oil sample af aniline in a test tube and heating the mixture until a homogeneous solution is obtained. Then, the tube is allowed to cool at a controlled rate. The temperature at which the oil and aniline phases separate out is recorded as the aniline point. A lower aniline point of oil means a higher percentage of aromatic hydrocarbons in it. Since aromatic hydrocarbons have a tendency to dissolve natural rubber and certain types of synthetic rubbers. Thus, higher the percentage of aromatic hydrocarbon or lower the aniline

ENGINEERING CHEMISTRY (BT-101) NOTES

point of a oil, more are the chances of deterioration of an oil when it comes in contact with rubber sealing's, packing, etc. Consequently, low aromatic content in the lubricants or their higher aniline point is desirable.

3. Steam Emulsion Number:

When water enters an oil system, as it often does in hot strip mills, turbulence caused by high volume flow results in the formation of emulsions and initiate mixture of oil and water. Depending on the base oil and the refining processes, some oils form emulsions with water more easily than others. Moreover, contamination also contribute to the formation of emulsions. These emulsions have poor lubricating properties thereby causing abrasion and wearing out of the lubricated parts of the machinery. The higher the percentage of water, the worse the lubricating properties. Hence, it is desirable that the lubricating oil should form such an emulsion with water which breaks off readily. This ability of lubricating oil to separate from water is called Demusibility. Oil that separates readily from water has good demulsibility. Oil that does not has poor demulsibility. The tendency of lubricant Water emulsion to break is determined by following *test*; steam at 100°C is bubbled through a test tube containing 20 mL of oil, till the increases to 90°C and the time is noted when the oil, and water separate out in distinct layers. The time in second in which oil and water emulsion separate out in distinct layers is called '**steam emulsion number**' (SEN) or '**emulsification number**'. The quicker the oil separates out from the emulsion, the lower the steam emulsion number and the better the lubricating oil for most purposes.

Significance: To avoid corrosion of polished steel surfaces like roll necks and to ensure proper lubrication, it is important to evaluate the speed of water and oil separation (demulsibility properties) of medium to high viscosity circulating oils used in rolling mills subject to cooling water contamination.

4. Neutralization Number

Lubricating Oil's acidity or alkalinity is determined in terms of neutralization number.

Neutralization number represents either the **Total Acid Number (TAN)**. "The number of milligrams of potassium hydroxide (KOH) needed to neutralize any acid in one gram of oil" or the **Total Base Number (TBN)** .The number of milligram of hydrochloric acid (HCl) needed to neutralize any base in one gram of oil. Determination of TAN is more common and its test procedure is given below:

Reagents:

- (i) 0.1 N alcoholic HCl solution,
- (ii) 0.1 N alcoholic KOH solution,
- (iii) Titration solvent: Made by mixing 500 mL of toluene and.5 mL of distilled water with 495mL of isopropyl alcohol. .
- (iv) Para-Naphtholbenzene indicator solution made by adding 1 g of dry indicator powder in 100 mL of isopropyl alcohol.

Test procedure:

- (i) Into 300 mL conical flask, take a weighed amount of the sample (20 gm sample for light-colored oil or 2 gm sample for dark-colored oil).
- (ii) Add 100 mL of titration solvent and 30 drops of indicator solution to the flask, then carefully swirl the mixture until the sample is completely dissolved.
- (iii) If the solution turns yellow-orange or deep orange in color, it means the oil sample was acidic [But if it assumes green or green-blue color, it means the oil sample was basic, then titration should be done with HCl].
- (iv) Slowly add the alc. KOH solution from the burette drop by drop with careful swirling until the green or green-blue end point is reached which persist for at least 15 s .The color change is reversed if talc . HCl is the titrating agent].
- (v) Read from the burette the number of mL of solution required to reach the end point. ,

Calculations: - Total mL of titrating solution / weight of sample used $\times 5.61$

Significance. This test shows relative changes in oil oil due to oxidation. Comparing the TAN or TEN with the values of new oil will indicate the development of harmful products or the effect of additive depletion. In fact, acid number greater than oil is usually taken as an indication of oxidation of the oil. This will consequently lead to corrosion, besides gum and sludge formation.

Saponification Number: **Saponification number** is defined as "The number of milligrams of potassium hydroxide required to saponify the fatty material present in one gram of the oil".

Determination. Saponification number is determined by refluxing a known quantity of oil with a known excess of potassium hydroxide solution and titrating the unused alkali against an acid.

Significance. Mineral oils do not undergo saponification but animal and vegetable oils undergo saponification. Hence, this test gives an indication of the amount of animal and vegetable oils added to mineral oils to improve oiliness. Moreover, most of the animal and vegetable oils possess their own characteristic saponification values. Any deviation from this value in a given sample indicates the probability and extent of adulteration. And last but not the least, this test helps us to ascertain whether the oil under study is animal/vegetable or mineral or compounded oil.

Iodine Value

The **iodine value** is defined as "the number of milligrams of iodine absorbed by one gram of the oil. It is determined by the addition of a measured, excess of a solution of iodine bromide in glacial acetic acid, to a weighed quantity of oil dissolved in CCl_4 or acetic acid. The addition of the halogenating agent to the double bond (s) is allowed to proceed to completion. The unused IBr is then back-titrated against standard hypo solution using starch as indicator.

Importance. The iodine value is a measure of the degree of unsaturation and hence the tendency of a fatty oil to absorb oxygen. Regardless of composition, any oil exposed to air and heat eventually combines with the oxygen in the air to form chemical compounds unsuitable for use as lubricants. Acids and gummy sludge are typical products of oxidation. Oils with high resistance to oxidation should be used in steam turbines and other large circulating oil systems, in which/oils attain high temperatures and remain in service for extended duration. Oils with lower oxidation resistance are satisfactory in application, where oil remains in service for periods or where makeup is high because of loses.

Carbon Residue

Lubricating oils contain high % of carbon in combined form. On being subjected to high temperatures, they decompose and form a carbonaceous deposit. There are two methods for measuring the amount of carbon residue or deposit remaining after a lubricating oil has been subjected to extreme heat.

(i) **The Conradson method.** It is conducted in the absence of air and is applicable for heavy residuals, crudes and non-volatile stock.

Procedure: A weighed amount of sample is placed in a silica crucible (of about 65-85 ml capacity), which in turn is put into skid more iron crucible having a close-fitting cover with a small horizontal opening .The crucibles are then placed into a larger third crucible also fitted with a cover fitted loosely to shaped iron hood. Heat is supplied from a Meker burner at certain prescribed rates till vapors of all volatile matter are burnt completely. After 30 minutes the silica crucible is removed, cooled in a desiccator, and weighed.

$$\% \text{ Carbon Residue} = \frac{\text{Weight of residue in crucible}}{\text{Weight of original oil sample}} \times 100$$

(ii) **The Ramsbottom method.** This method is used with the more fluid products.

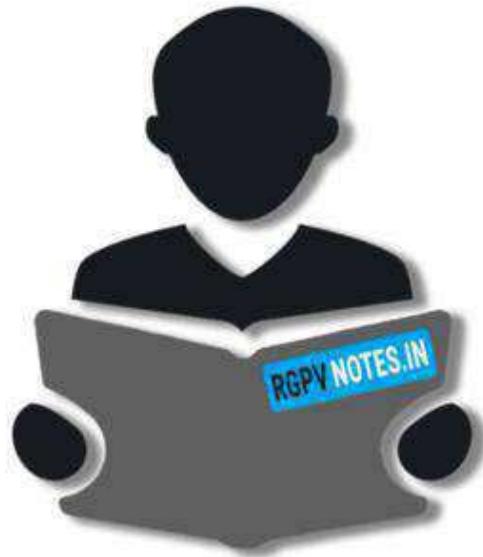
Procedure. A weighed sample is placed in a special glass bulb with capillary opening. (The oil is injected into the bulb through the capillary inlet by means of a syringe). This glass bulb containing

ENGINEERING CHEMISTRY (BT-101) NOTES

sample is inserted in one of the holes of an electrically heated small furnace (Ramsbottom apparatus) maintained at approximately 550°C. At this temperature, all volatile matter escapes the bulb capillary leaving a residue that undergoes cracking and possible coke formation. After the heating, the bulb is taken out, cooled in a desiccator and weighed.

% Carbon Residue = Weight of residue in crucible / Weight of original oil sample × 100

Significance of carbon-residue tests. Certain lubricating oils tend to deposit carbon in the combustion chambers of internal combustion engines, due to the carbonizing of the lubricating oil carried up past the piston rings into the combustion chambers. Incomplete combustion of fuel also results in carbon deposition. Excessive buildup of carbon deposits in the combustion chamber results in decreased volume of the charge at the end of the compression stroke giving increasing compression ratio which eventually leads to detonation. The tests are used in the evaluation of base crudes and feed stocks for the formulation of lubricants and fuels to be used in extreme temperature service.



RGPVNOTES.IN

We hope you find these notes useful.

You can get previous year question papers at
<https://qp.rgpvnotes.in> .

If you have any queries or you want to submit your
study notes please write us at
rgpvnotes.in@gmail.com



LIKE & FOLLOW US ON FACEBOOK
facebook.com/rgpvnotes.in



RGPVNOTES.IN

Program : **B.Tech**

Subject Name: **Engineering Graphics**

Subject Code: **BT-105**

Semester: **1st**



LIKE & FOLLOW US ON FACEBOOK
facebook.com/rgpvnotes.in

Module - III

Projections of Regular Solids covering, those inclined to both the Planes- Auxiliary Views; Draw simple annotation, dimensioning and scale. Floor plans that include: windows, doors, and fixtures such as WC, bath, sink, shower, etc.

Solid:

A solid has three dimensions, the length, breadth and thickness or height. A solid may be represented by orthographic views, the number of which depends on the type of solid and its orientation with respect to the planes of projection. Solids are classified into two major groups. (i) Polyhedra and (ii) Solids of revolution

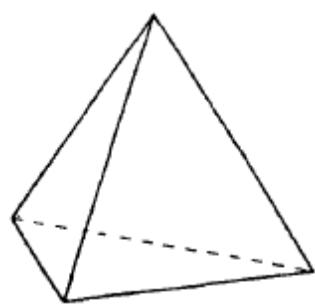
Classification of Solids:**1. Polyhedra**

Polyhedra are defined as a solid bounded by plane surfaces called faces. They are:

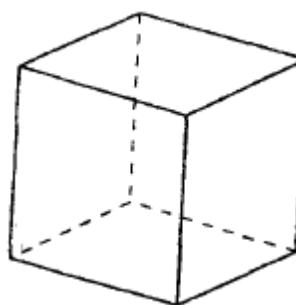
- (i) Regular polyhedra
- (ii) Prisms and
- (iii) Pyramids

(i) Regular Polyhedra

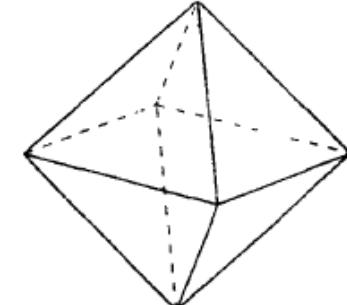
A polyhedron is said to be regular if its surfaces are regular polygons. The following are some of the regular polyhedra.



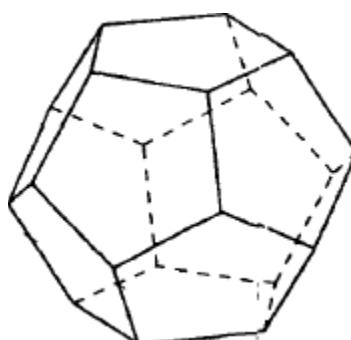
(a) Tetrahedron



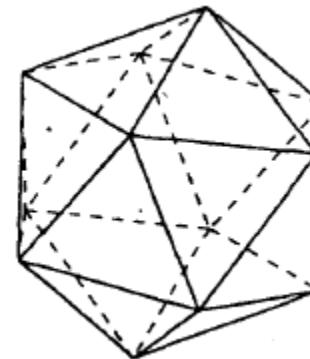
(b) Hexahedron (cube)



(c) Octahedron



(d) Dodecahedron



(e) Icosahedron

Fig. 3.1

- (a) Tetrahedron: It consists of four equal faces, each one being a equilateral triangle.
- (b) Hexahedron (cube): It consists of six equal faces, each a square.
- (c) Octahedron: It has eight equal faces, each an equilateral triangle.
- (d) Dodecahedron: It has twelve regular and equal pentagonal faces.
- (e) Icosahedron: It has twenty equal, equilateral triangular faces.

(ii) Prisms

A prism is a polyhedron having two equal ends called the bases parallel to each other. The two bases are joined by faces, which are rectangular in shape. The imaginary line passing through the centers of the bases is called the axis of the prism.

A prism is named after the shape of its base. For example, a prism with square base is called a square prism, the one with a pentagonal base is called a pentagonal prism, and so on (Fig. 3.2) The nomenclature of the prism is given in Fig. 3.3.

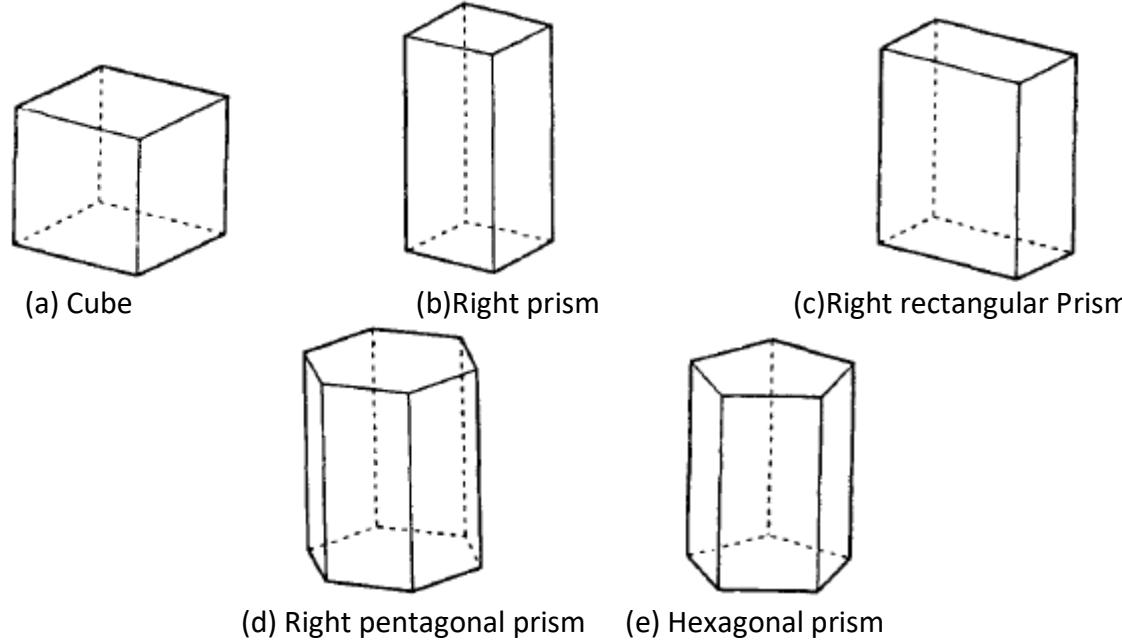


Fig. 3.2

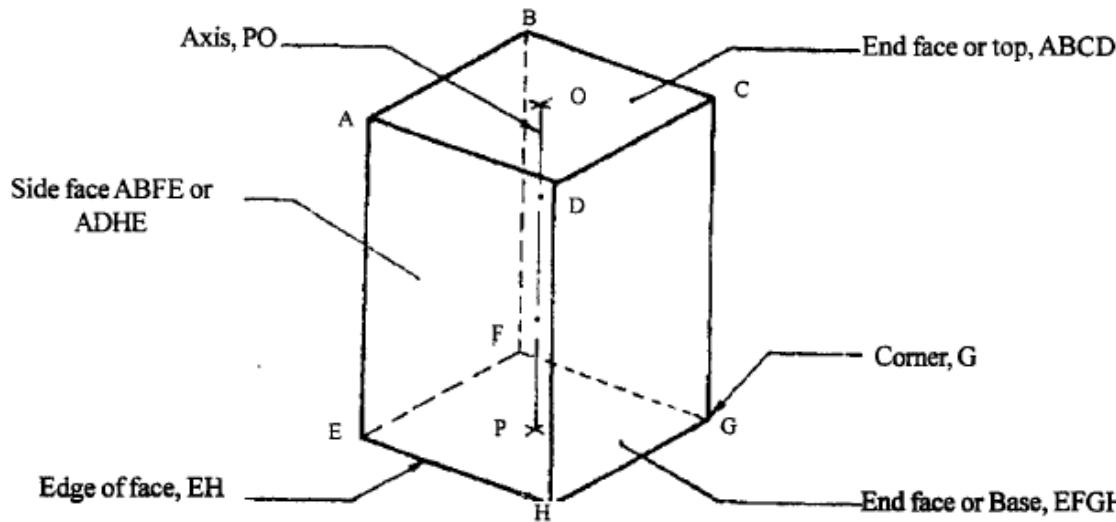


Fig. 3.3 Nomenclature of a Square Prism

(iii) Pyramids

A pyramid is a polyhedron having one base, with a number of isosceles triangular faces, meeting at a point called the apex. The imaginary line passing through the centre of the base and the apex is called the axis of the pyramid.

The pyramid is named after the shape of the base. Thus, a square pyramid has a square base and pentagonal pyramid has pentagonal base and so on (Fig. 3.4). The nomenclature of a pyramid is shown in Fig. 3.5.

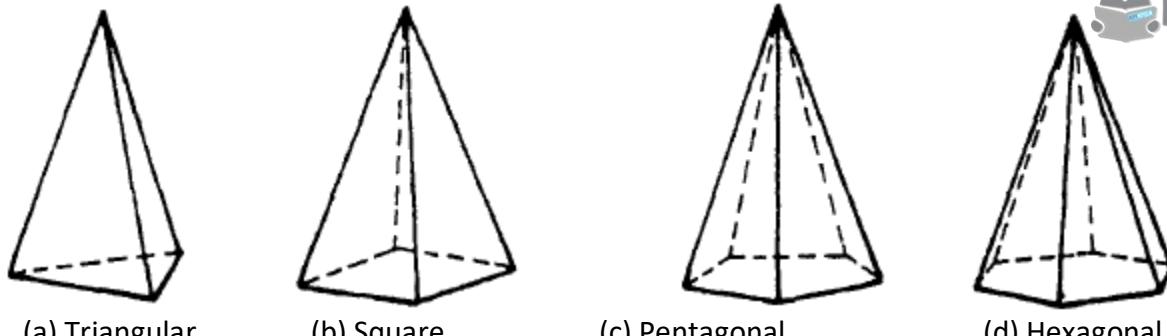


Fig. 3.4

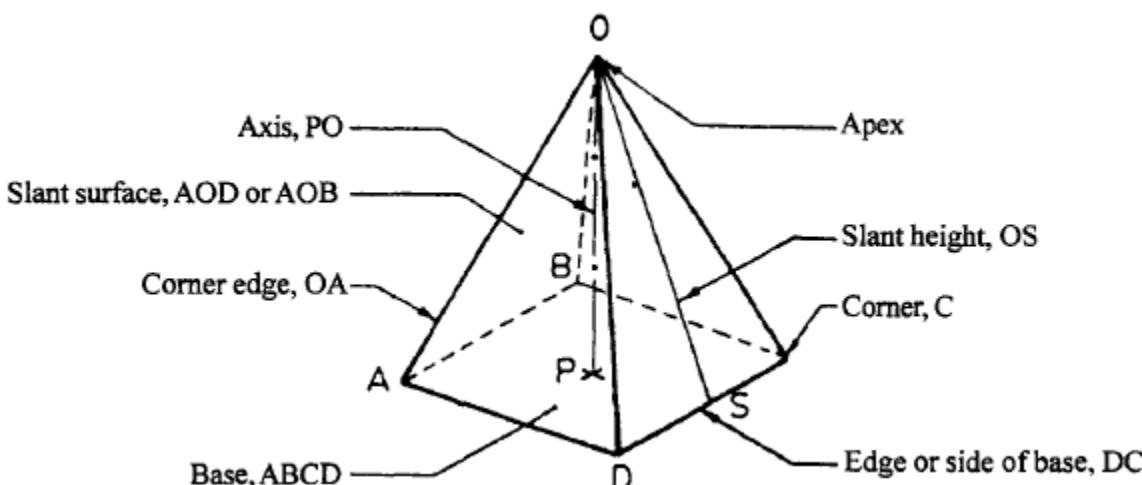


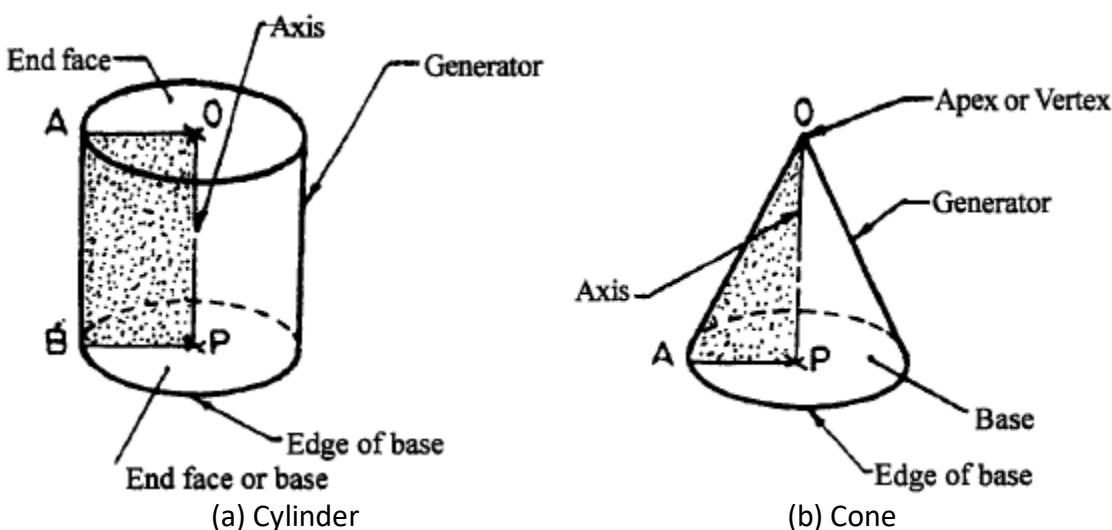
Fig. 3.5 Nomenclature of a Square Pyramid

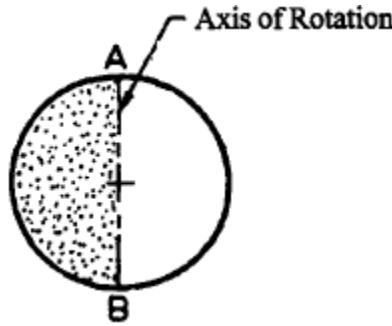
2. Solids of Revolution

If a plane surface is revolved about one of its edges, the solid generated is called a solid of revolution. The examples are (i) Cylinder, (ii) Cone, (iii) Sphere

Frustums and Truncated Solids

If a cone or pyramid is cut by a section plane parallel to its base and the portion containing the apex or vertex is removed, the remaining portion is called frustum of a cone or pyramid





(c) Sphere

Fig. 3.6 Solid of Revolution

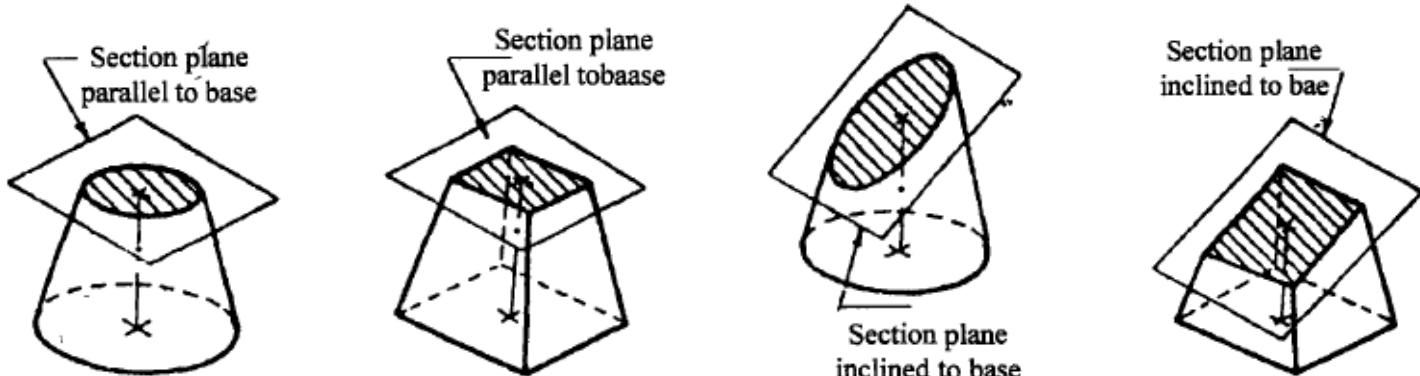


Fig. 3.7 Frustum of a Solid and Truncated Solids

Prisms (problem) Position of a Solid with Respect to the Reference Planes

The position of solid in space may be specified by the location of the axis, base, edge, diagonal or face with the principal planes of projection. The following are the positions of a solid considered.

1. Axis perpendicular to one of the principal planes.
2. Axis parallel to both the principal planes.
3. Axis inclined to one of the principal planes and parallel to the other.
4. Axis inclined to both the principal planes.

The position of solid with reference to the principal planes may also be grouped as follows:

1. Solid resting on its base.
2. Solid resting on anyone of its faces, edges of faces, edges of base, generators, slant edges, etc.
3. Solid suspended freely from one of its corners, etc.

1. Axis perpendicular to one of the principal planes

A) Axis perpendicular to H. P.

When the axis of a solid is perpendicular to one of the planes, it is parallel to the other. Also, the projection of the solid on that plane will show the true shape of the base.

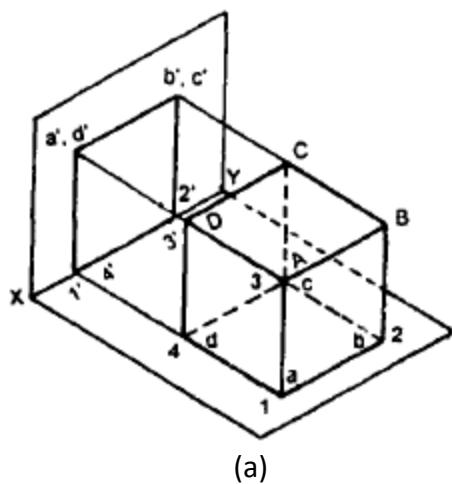
When the axis of a solid is perpendicular to H.P., the top view must be drawn first and then the front view is projected from it. Similarly when the axis of the solid is perpendicular to V.P., the front view must be drawn first and then the top view is projected from it.

Problem: Draw the projections of a cube of 35 mm side, resting on one of its faces (bases) on H.P., such that one of its vertical faces is parallel to and 10 mm in front of V.P.

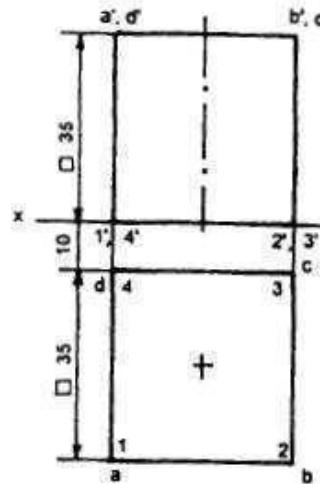
Construction: Above figure shows the cube positioned in the first quadrant.

1. Draw the top view such that one of its edges is 10mm below XY.
2. Obtain the front view by projection, keeping one of its bases on XY.

Note: (i) For the cube considered ABCD is the top base and 1234 the bottom base, (ii) Fig. 3.8 c shows the projections of a cube, resting on one of its bases on H.P. such that an edge of its base is inclined at 30° to V.P.



(a)



(b)

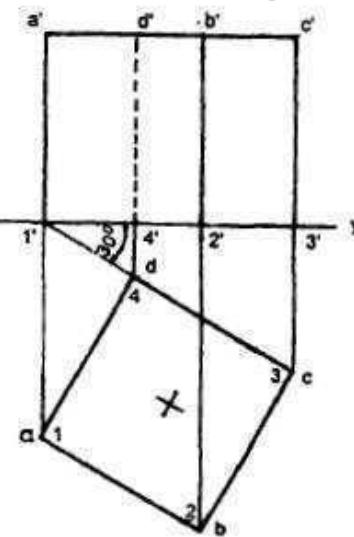


Fig. 3.8

B) Axis perpendicular to V. P.

Problem: A square prism with side of base 35 mm and axis 50 mm long, lies with one of its longest edges on H.P such that its axis is perpendicular to V.P. Draw the projections of the prism when one of its rectangular faces containing the above longer edge is inclined at 30° to H.P.,

Construction:

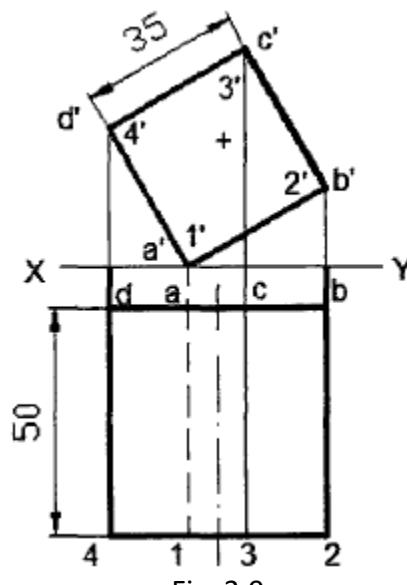


Fig. 3.9

1. Draw the front view which is a square of 35mm such that one of its corners is on XY and a side passing through it is making 30° with XY.

2. Obtain top view by projection, keeping the length as 50mm.

Note: The distance of the base nearer to V.P is not given in the problem. Hence, the top view may be drawn keeping the base nearer to XY at any convenient distance.

2. Axis parallel to both the principal planes

When the axis of solid is parallel to both the planes, neither the front view nor the top view reveal the true shape of the base. In such case, the side view must be drawn first which shows the true shape of the base. The front and top view are then projected from the side view.

Problem: Hexagonal prism with side of base 25 mm and axis 60 mm long is lying on one of its rectangular faces on HP. Draw the projections of the prism when its axis is parallel to both HP and V.P.

Construction:

1. Draw the right side view of the hexagon, keeping an edge on XY.
2. Draw the second reference line X_1Y_1 perpendicular to XY and to the right of the above view at any convenient location.
3. Obtain the front view by projection, keeping its length equal to 60 mm.
4. Obtain the top view by projecting the above views.

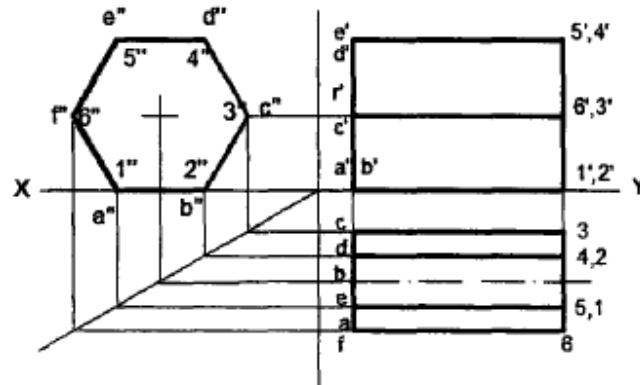


Fig. 3.10

3. Axis inclined to one of the principal planes and parallel to the other:

A) Axis inclined to H.P. and parallel to V.P.

When the axis of a solid is inclined to any plane, the projections are obtained in two stages. In the first stage, the axis of the solid is assumed to be perpendicular to the plane to which it is actually inclined and the projections are drawn. In second stage, the position of one of the projections is altered to satisfy the given condition and the other view is projected from it. This method of obtaining the projections is known as the change of position method.

Problem: A pentagonal prism with side of base 30 mm and axis 60 mm long is resting with an edge of its base on HP, such that the rectangular face containing that edge is inclined at 60° to HP. Draw the projections of the prism when its axis is parallel to V.P.

Construction:

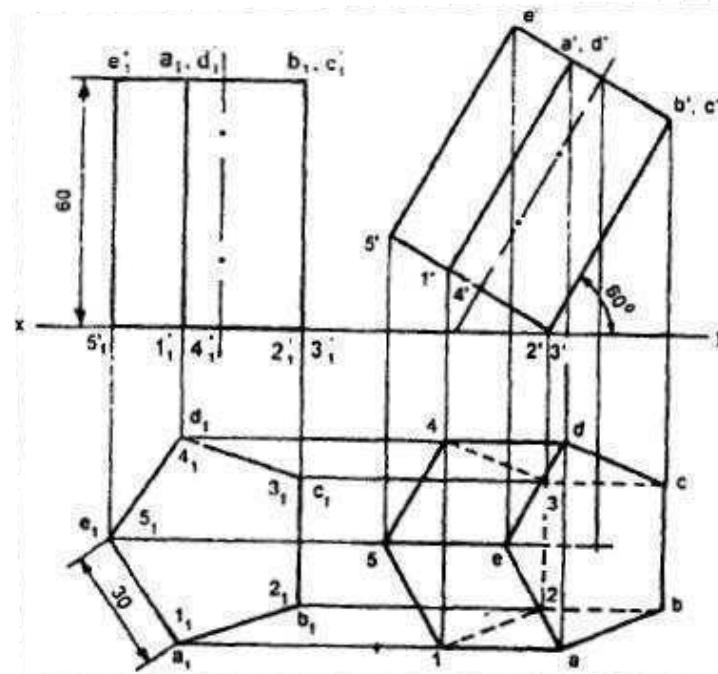


Fig. 3.11

Stage 1

Assume that the axis is perpendicular to H.P.

1. Draw the projections of the prism keeping an edge of its base perpendicular to V.P.

Stage 2

1. Rotate the front view so that the face containing the above edge makes the given angle with the H.P.
2. Redraw the front view such that the face containing the above edge makes 60° with XY. This is the final front view.

3. Obtain the final top view by projection.

Note: For completing the final projections of the solids inclined to one or both the principal planes, the following rules and sequence may be observed.

- (i) Draw the edges of the visible base. The base is further away from XY in one view will be fully visible in the other view.
- (ii) Draw the lines corresponding to the longer edges of the solid, keeping in mind that the lines passing through the visible base are invisible.
- (iii) Draw the edges of the other base.

B) Axis inclined to V.P. and parallel to H.P.

Problem: A pentagonal prism with side of base 25 mm and axis 50 mm long lies on one of its faces on H.P., such that its axis is inclined at 45° to V.P. Draw the projections.

Construction:

1. Assuming that the axis is perpendicular to V.P, draw the projections keeping one side of the pentagon coinciding with XY.
2. Redraw the top view so that the axis is inclined at 45° to XY. This is the final top view.
3. Obtain the final front view by projection.

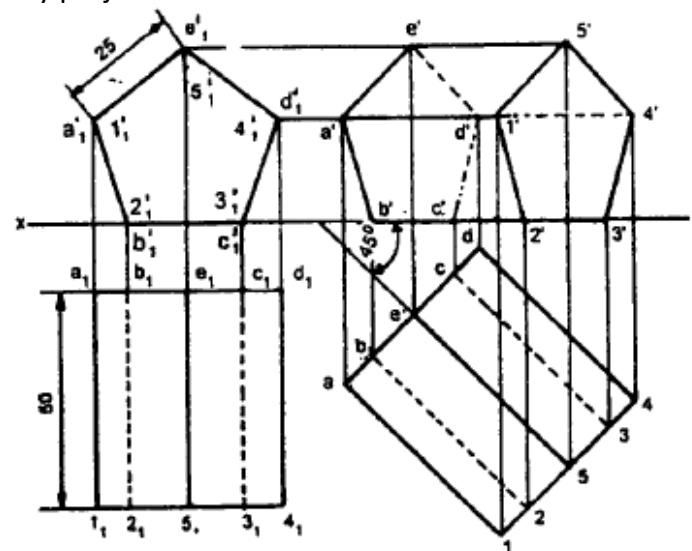


Fig. 3.12

4. Axis inclined to both the principal planes

A solid is said to be inclined to both the planes when (i) the axis is inclined to both the planes, (ii) the axis is inclined to one plane and an edge of the base is inclined to the other. In this case the projections are obtained in three stages.

Stage I

Assume that the axis is perpendicular to one of the planes and draw the projections.

Stage II

Rotate one of the projections till the axis is inclined at the given angle and project the other view from it.

State III

Rotate one of the projections obtained in Stage II, satisfying the remaining condition and project the other view from it.

Problem: A square prism with side of base 30 mm and axis 50 mm long has its axis inclined at 60° to H.P., on one of the edges of the base which is inclined at 45° to V.P.

Construction:

1. Draw the projections of the prism assuming it to be resting on one of its bases on H.P. with an edge of it perpendicular to V.P.
2. Redraw the front view such that the axis makes 60° with XY and project the top view from it.
3. Redraw the top view such that the edge on which the prism is resting on H.P. is inclined at 45° to XY. This is the final top view.
4. Obtain the final front view by projection.

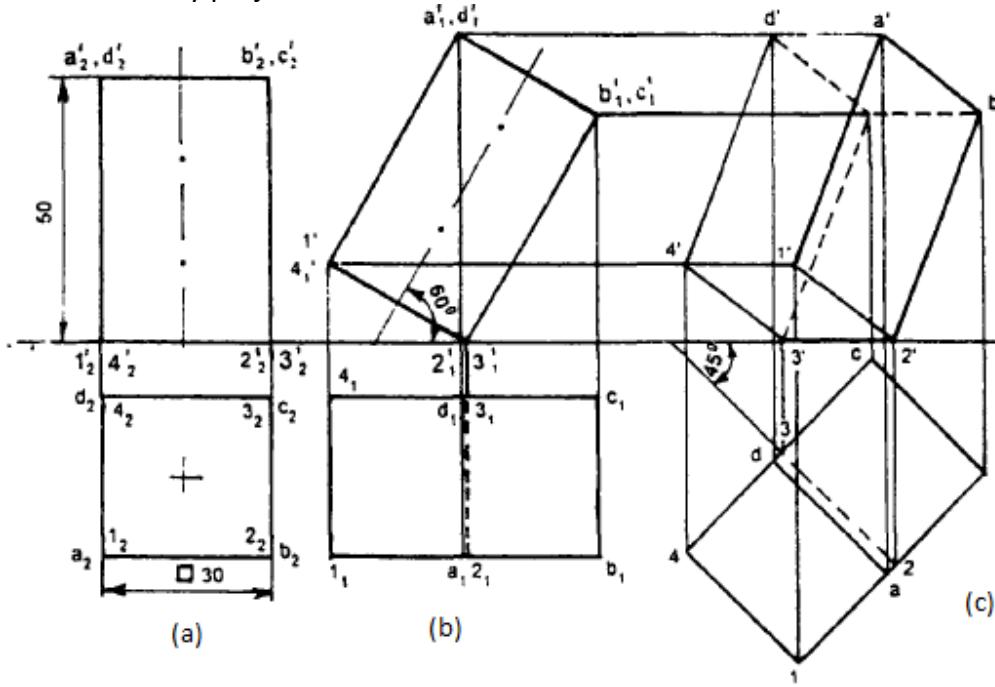


Fig. 3.13

Pyramids**1. Axis perpendicular to one of the principal planes****A) Axis perpendicular to H. P.**

Problem: A square pyramid with side of base 30 mm and axis 50 mm long is resting with its base on HP. Draw the projections of the pyramid when one of its base edges is parallel to V.P. The axis of the pyramid is 30 mm in front of V.P.

Construction:

1. Draw the top view, a square, keeping its centre at 30mm from XY and with an edge parallel to XY.
2. Obtain the front view by projection keeping the height equal to 50 mm and the base lying on XY.

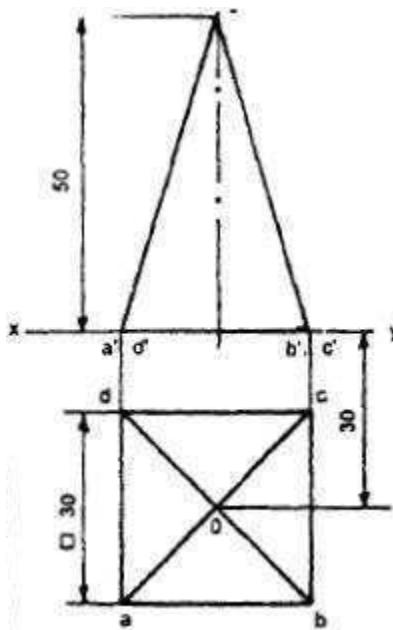


Fig. 3.14

B) Axis perpendicular to V. P.

Problem: Draw the projections of a pentagonal pyramid of side of base 30mm and axis 50mm long when its axis is perpendicular to V.P. and an edge of its base is perpendicular to H.P

Construction:

1. Draw the front view of the pyramid which is a pentagon, keeping one of its side perpendiculars to XY.
2. Obtain the top view by projection keeping the axis length equal to 50 mm.

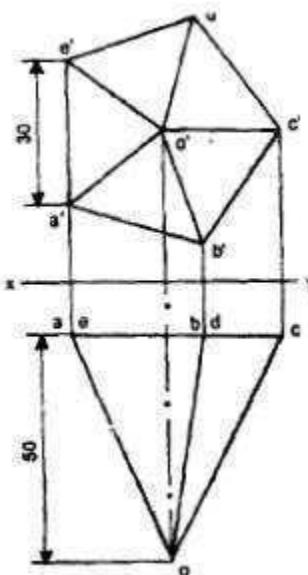


Fig. 3.15

2. Axis parallel to both the principal planes

Problem: A pentagonal pyramid with side of base 30 mm and axis 60 mm long rests with an edge of its base on HP such that its axis is parallel to both HP and V.P. Draw the projection of the solid.

Construction:

1. Draw the projections of the pyramid with its base on H.P and an edge of the base (BC) perpendicular to V.P.
2. Redraw the front view such that b(c) lies on XY and the axis is parallel to XY which is the final front view.
3. Obtain the final top view by projection.

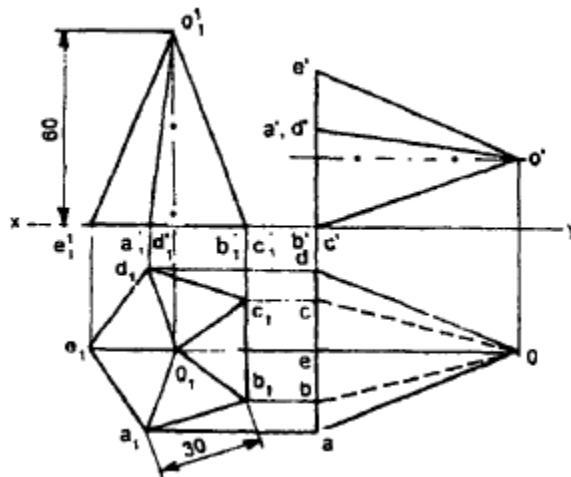


Fig. 3.16

3. Axis inclined to one of the principal planes and parallel to the other:

Problem: A pentagonal pyramid with side of base 25 and axis 50mm long is resting on one of its faces on HP such that its axis is parallel to V.P. Draw the projections.

Construction:

1. Assuming the axis is perpendicular to H.P draw the projections keeping one edge of the base perpendicular to V.P.
2. Redraw the front view so that the line $o'-c$ (d) representing the slant face, coincides with XY. This is the final front view.
3. Obtain the final top view by projection.

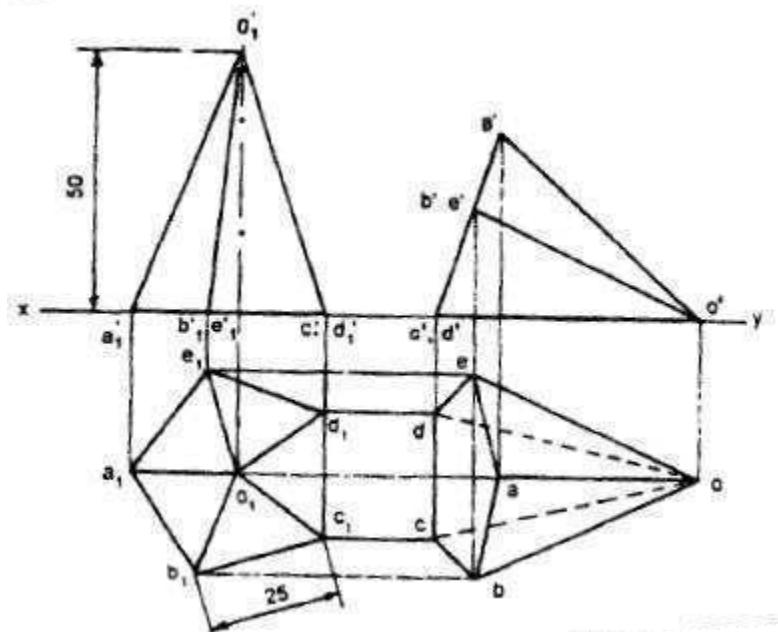


Fig. 3.17

Cone and cylinder

Problem: Draw the projection of a cone of base 40 mm diameter, axis 60 mm long when it is resting with its base on H.P.

Construction:

1. Draw the reference line XY and locate O at a convenient distance below it.
2. With centre O and radius 20 mm draw a circle forming the top view.
3. Obtain the front view by projection, keeping the height equal to 60 mm and the base coinciding with XY.

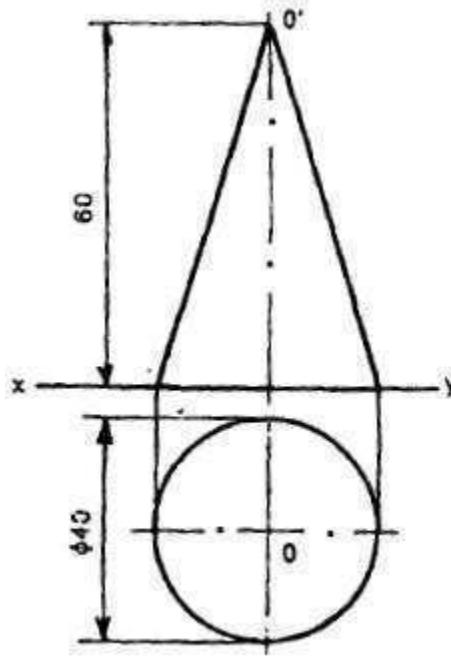


Fig. 3.18

Problem: A cone with base 30 mm diameter and axis 45 mm long lies on a point of its base on V.P. such that the axis makes an angle 45° with V.P. Draw the projections of the cone.

Construction:

1. Draw the projections of the cone assuming that the cone is resting with its base on V.P.
2. Divide the circle into a number of equal parts and draw the corresponding generators in the top view.
3. Redraw the top view so that the axis makes 45° with XY. This is the final top view.
4. Obtain the final view by projection.

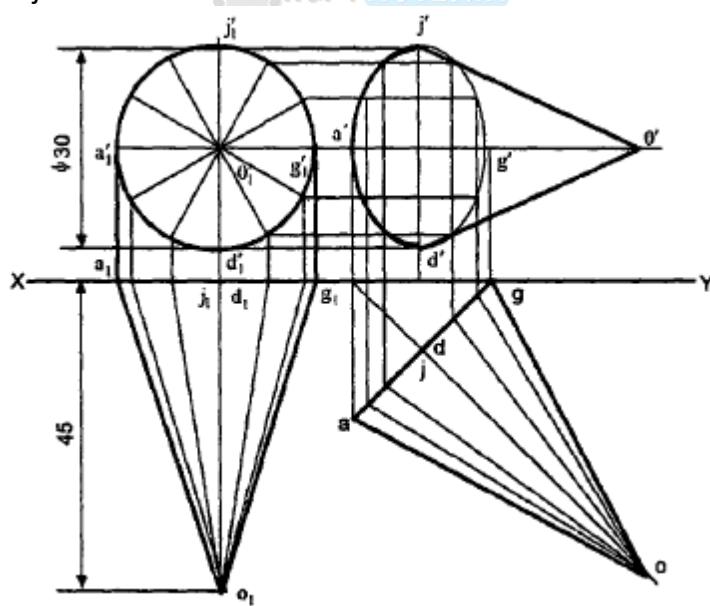


Fig. 3.19

Problem: A cylinder with base 40 mm diameter and 50 mm long rests on a point of its base on HP such that the axis makes an angle of 30° with HP. Draw the projections of the cylinder.

Construction:

1. Draw the projection of the cylinder assuming that the cylinder is resting with its base on H.P.
2. Divide the circle into a number of equal parts and obtain the corresponding generators in the front view.
3. Redraw the front view such that its axis makes 30° with XY. This is the final front view.
4. Obtain the final top view by projection.

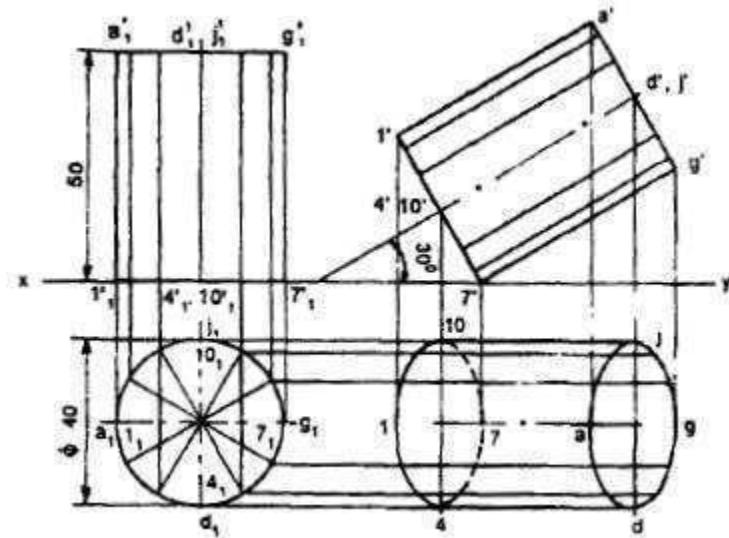
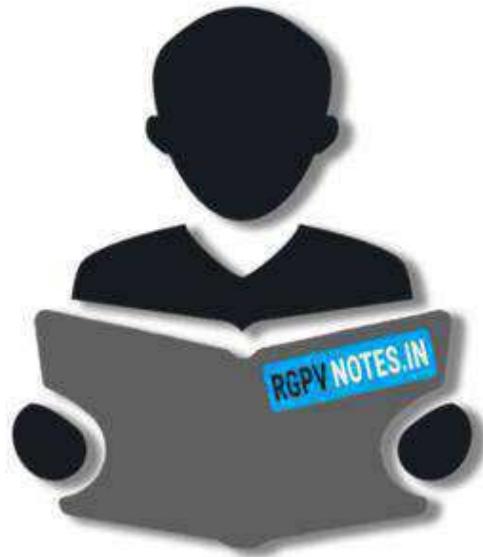


Fig. 3.20



RGPVNOTES.IN

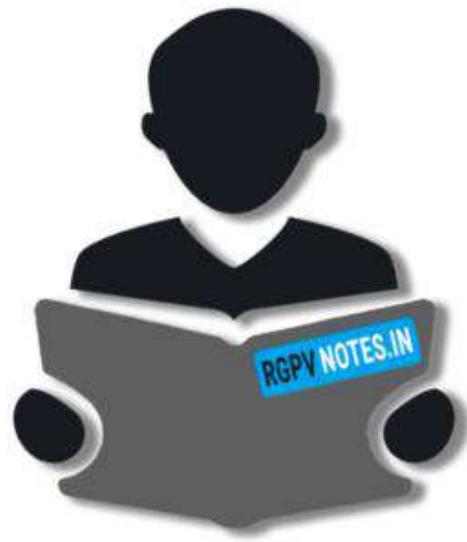
We hope you find these notes useful.

You can get previous year question papers at
<https://qp.rgpvnotes.in> .

If you have any queries or you want to submit your
study notes please write us at
rgpvnotes.in@gmail.com



LIKE & FOLLOW US ON FACEBOOK
facebook.com/rgpvnotes.in



RGPVNOTES.IN

Program : **B.Tech**

Subject Name: **Basic Electrical & Electronics Engineering**

Subject Code: **BT-104**

Year: **1st**



LIKE & FOLLOW US ON FACEBOOK
facebook.com/rgpvnotes.in

BASIC ELECTRICAL & ELECTRONICS ENGINEERING NOTES

Unit - IV Electrical Machines - Syllabus:

Construction, Working Principle, Classification and Application of DC machine, Induction machine and Synchronous machine. Working principle of 3-Phase induction motor, Concept of slip in 3-Phase induction motor, Explanation of Torque-slip characteristics of 3-Phase induction motor. Types of losses occurring in electrical machines.

D.C. MACHINE

A DC machine can work as a motor as well as a generator. There is no constructional difference between a DC motor and a DC generator.

Constructional Features: Fig. - 1 shows the constructional details of a DC Machine. Some of the essential parts of a DC machine are Yoke, Main poles, Field coil, Armature core, Armature winding, Commutator, Brushes, Bearings and End Covers.

Yoke: It is the outermost covering made of cast iron which provides support for the main poles. It also carries the magnetic flux produced by the poles.

Main poles: They are made up of laminations bunched together and fixed to the yoke. The pole core accommodates the field coils. The pole shoe helps in spreading the flux in the air gap.

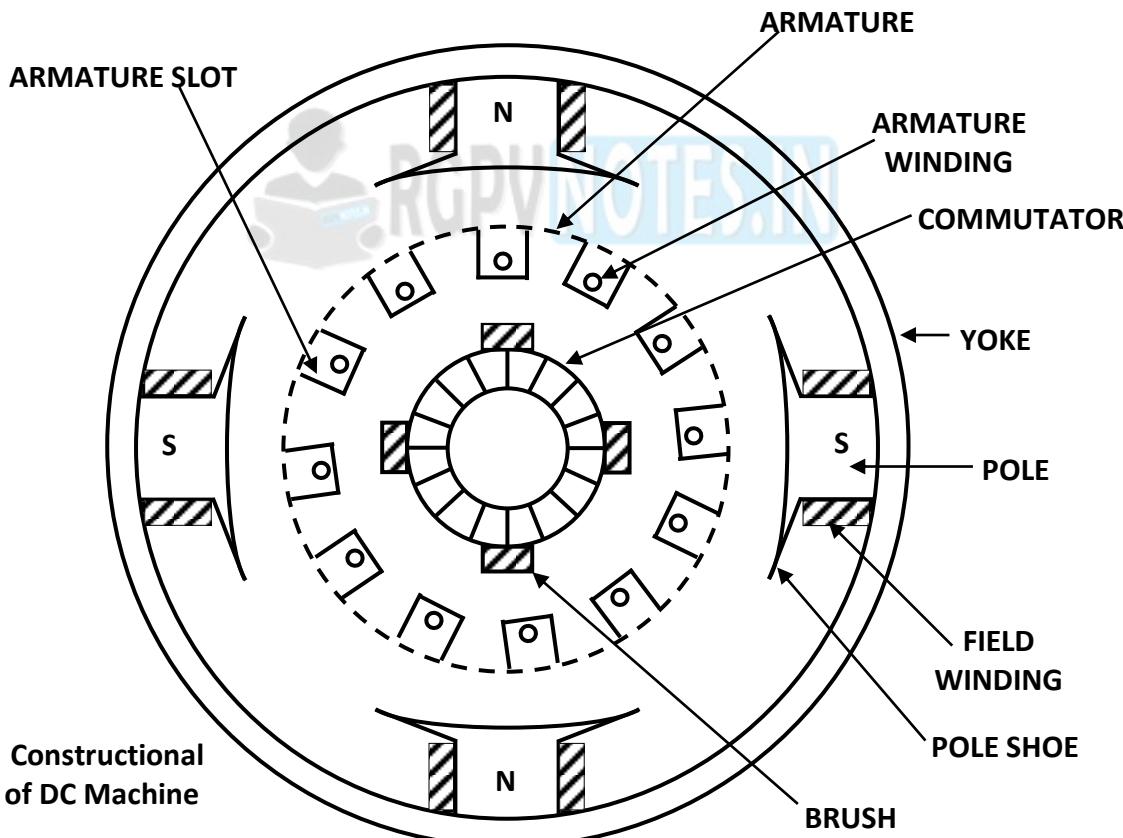


Fig. - 1 Constructional details of DC Machine

Field coil: They are placed around the pole core supported by the pole shoe. The field coils after being mounted on the main poles are connected such that, when a DC exciting current flows through them, the main poles alternately become north and south poles producing the necessary flux.

BASIC ELECTRICAL & ELECTRONICS ENGINEERING NOTES

Armature core: It is a laminated cylindrical core keyed to the machine shaft with slots on the outer periphery to accommodate the armature winding as shown in Fig. - 2. The important function of the core is to provide a path of very low reluctance to the flux through the armature from North pole to South pole.

Armature winding: Copper wire is used to make the armature winding in the slots of the armature core, which form the conductors. The conductors are insulated from each other as well as from the core. The armature winding of a DC machine forms a closed circuit. Depending on the manner in which the conductors are connected to the commutator segments, we have lap winding and wave winding.

Commutator: It converts the alternating current induced in the armature conductors into unidirectional current in the external load circuit. It is a cylindrical structure mounted on the shaft of the armature core on one side. It has many segments of copper that are insulated from each other by mica sheets. The number of segments will be equal to the number of armature coils. The commutator facilitates for the collection of current from the armature conductors.

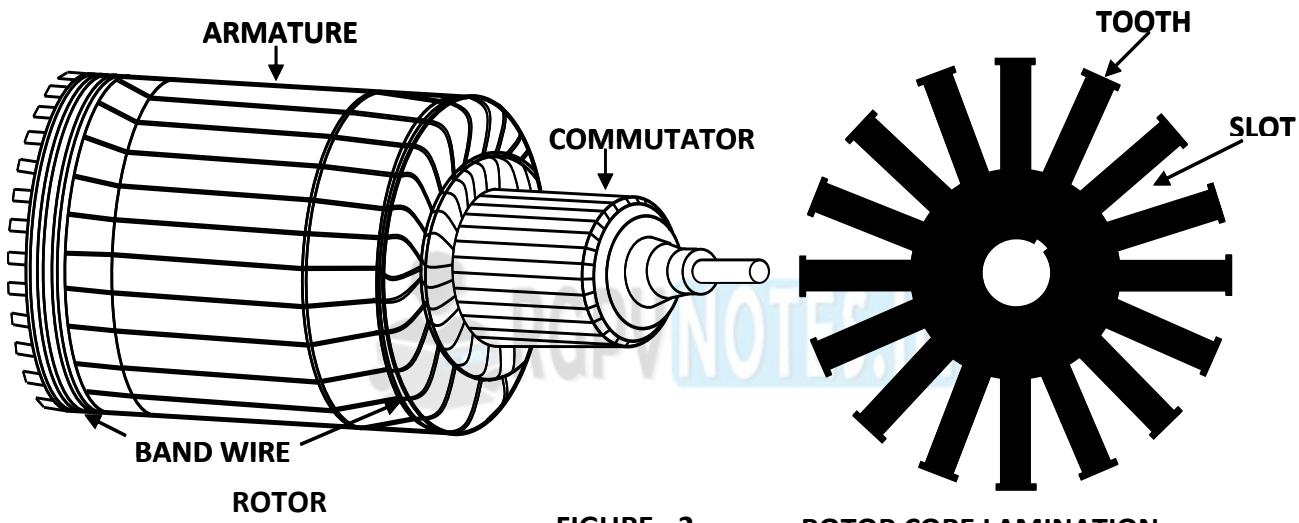


FIGURE - 2

ROTOR CORE LAMINATION

Brushes: They are made of carbon and will be equal to the number of main poles. They rest on the commutator and help in collection of current from the commutator. The brushes are placed in brush holders, which are accommodated inside the end covers.

Bearings: The bearings are fixed in the end covers. The shaft of the armature core is held on either side by the bearings. Their function is to reduce friction.

End Covers: They cover the yoke of the machine on either side. They are made of thick sheet metal. They accommodate bearings and brush holders. They also provide ventilation.

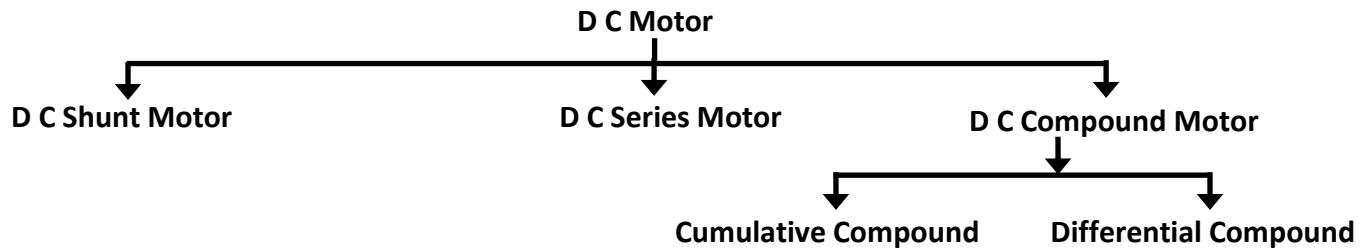
D.C. MOTOR

Working principle of DC machine as motor: A DC motor converts electrical energy into mechanical energy. Its operation is based on the principle that when a current carrying conductor is placed in a magnetic field, the conductors experience a mechanical force, whose direction is given by Fleming's left hand rule and its magnitude is given by $F = Bil$ Newtons

BASIC ELECTRICAL & ELECTRONICS ENGINEERING NOTES

Working as a motor: The field winding is excited such that the main poles become alternately north and south, the magnetic field is established in the air gap between the main poles and armature core. The DC supply is also given to the armature terminals so that current flows through the armature conductors. The current carrying conductors when placed in a magnetic field experience a mechanical force. The direction of the force developed in all the conductors will be in the same direction. These forces acting on the shaft give rise to torque.

Types of DC Motor:



Applications of DC Motors: Shunt motors are used for applications which require medium level torque such as Blowers, Fans, Centrifugal and Reciprocating pumps, Lathe machines, Machine tools.

Series motors are used for applications which require high starting torque such as in Cranes, Trolleys, Conveyers, Elevators and Electric locomotives.

Cumulative Compound motors are also used for applications which require high starting torque such as in Rolling mills, Heavy Planers and Elevators.

Differential Compound motors are not suitable for practical applications.

D.C. GENERATOR

Working principle of DC machine as generator: A DC generator is a machine that converts mechanical energy into electrical energy. The energy conversion in a DC generator is based on Faradays laws of electromagnetic induction i.e. the principle of production of dynamically induced emf. Whenever a conductor cuts magnetic flux an emf is induced, which will cause a current to flow, if the conductor circuit is closed. The direction of induced emf and hence current is given by Fleming's right hand rule. Hence, the basic requirement for a generator will be a magnetic field, some number of conductors and motion of the conductors with respect to the magnetic field.

Working as a generator: When DC supply is given to the field coils, the main poles get magnetized and the pole shoes spread the magnetic flux through the air gap over the armature conductors accommodated in the armature core. If the armature core is rotated by a prime mover, armature conductors cut the magnetic flux and have emf induced in them. The armature coils connected to the commutator segments form a

BASIC ELECTRICAL & ELECTRONICS ENGINEERING NOTES

closed loop. The brushes moving over the commutator segments collect the current. If the brushes are externally connected to a load circuit, current flows in the external circuit.

Emf equation of DC generator:

Let Φ = Flux per pole in weber ; Z = Total number of armature conductors ; P = Number of poles

N = Speed of armature in rpm ; A = Number of parallel paths, where $A = 2$ for wave winding
and $A = P$ (i.e. No. of poles) for lap winding

Eg = emf of the generator = emf per parallel path

Flux cut by one conductor in one revolution of the armature, $d\Phi = P\Phi$ weber

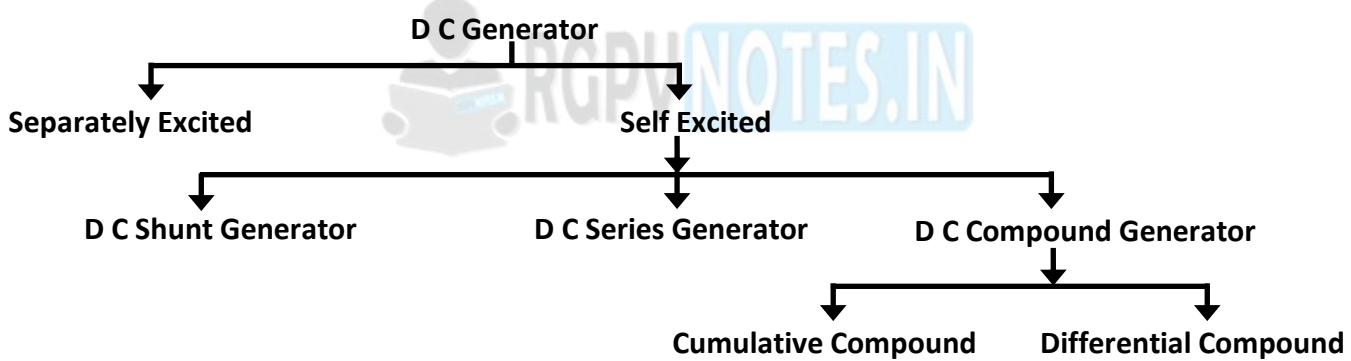
Time taken to complete one revolution, $dt = 60/N$ seconds

$$\text{emf generated per conductor} = \frac{d\Phi}{dt} = \frac{P\Phi}{60/N} = \frac{P\Phi N}{60} \text{ volts}$$

emf of generator, $E_g = \text{emf per parallel path}$

$$= \text{emf per conductor} \times \text{No. of conductors in series per parallel path}$$

$$\therefore E_g = \frac{P\Phi N}{60} \times \frac{Z}{A} = \frac{P\Phi Z N}{60 A} \text{ volts}$$

Types of DC Generator:

Applications of DC Generators: Separately excited generators are used for feeding supply to equipment used for Electro-plating and Electro-refining of materials.

Shunt generators feed supply to equipment used for Battery charging and Lighting loads.

Series generators are used as Boosters on DC feeders and for applications that need constant current.

Cumulative compound generators are usually used for Lighting purposes.

Differential compound generators are used for feeding supply to Electric arc welding equipment.

THREE PHASE INDUCTION MOTOR

An induction motor is an AC machine that converts electrical energy into mechanical energy. The rotor of the Induction motor gets its excitation through induction hence it is called Induction motor.

Production of Rotating magnetic field: Let us consider the 3 windings of the three phases of an induction motor stator with assumed positive directions of currents and the mmf space phasors as shown in Fig. -3.

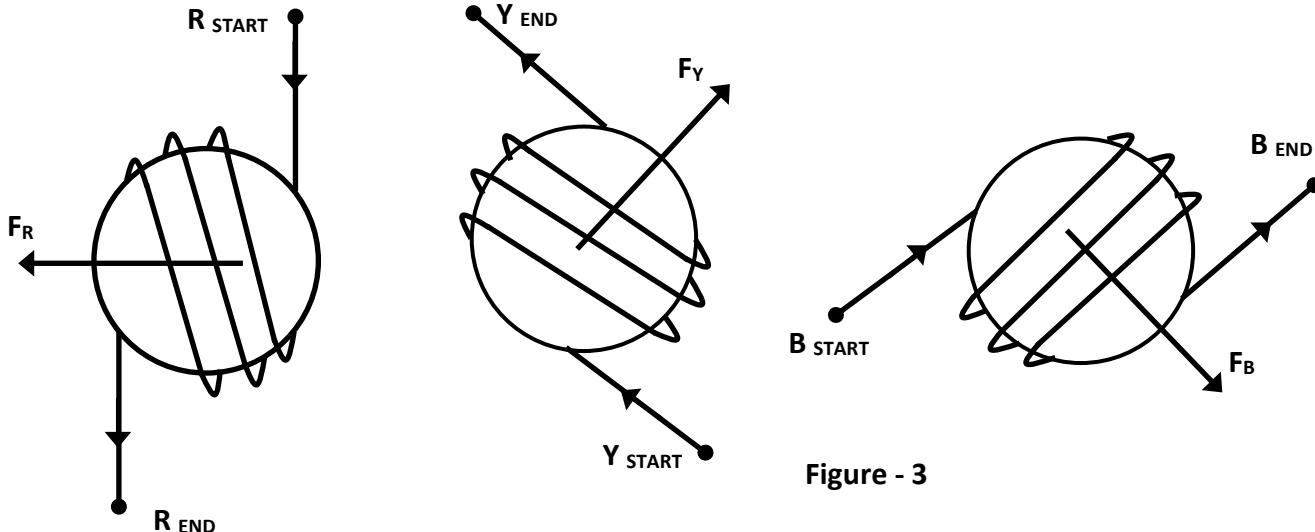


Figure - 3

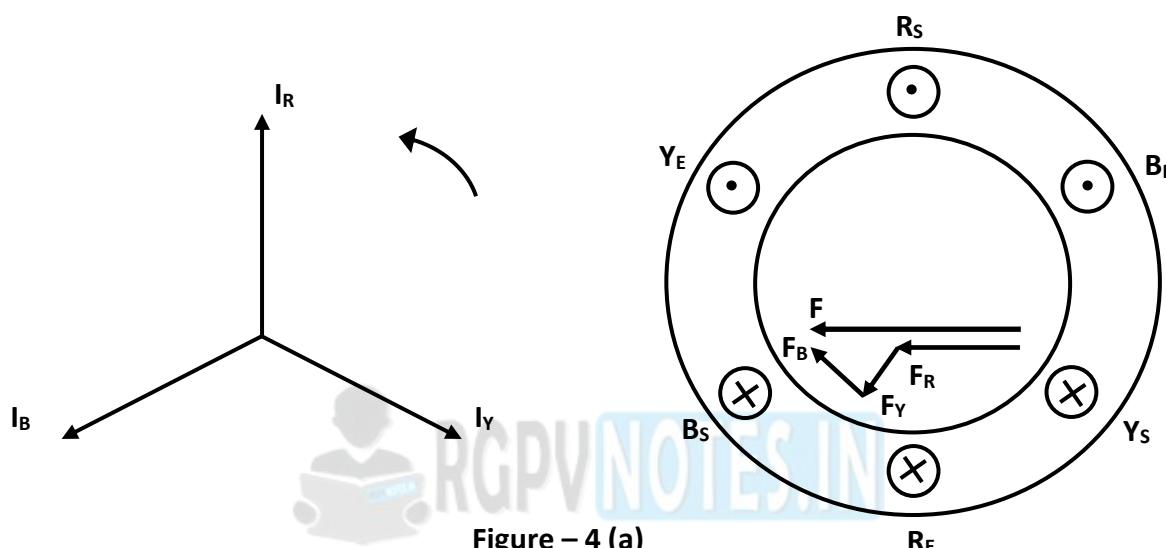


Figure - 4 (a)

The winding of phase Y is displaced from the winding of phase R by 120° and the winding of phase B is displaced from the winding of phase Y by 120° . F_R , F_Y and F_B denote the mmf of the three windings at the instant when currents in them are at their positive maximum values. As the currents alternate, the individual mmf will vary in magnitude and change direction so as to follow the changes in the currents. Let us consider two instants as shown in Fig -4.

Figure - 4(a) corresponds to the instant when I_R is at its positive maximum value, I_Y is negative and half its maximum value, I_B is also negative and half its maximum value. So F_Y and F_B have half the magnitude of F_R since currents in phases Y and B are negative, F_Y and F_B are shown opposite to that shown in Figure - 3, F is the resultant of F_R , F_Y and F_B .

Fig 4(b) corresponds to the instant 30° later when I_R is positive and $\sqrt{3}/2$ of its maximum value, I_Y is zero, I_B is negative and $\sqrt{3}/2$ of its maximum value. So F_Y is zero, F_R and F_B are equal in magnitude. It is observed that the resultant mmf, F has the same magnitude but has advanced by 30° . So an elapse of 30 electrical degrees in time results in a rotation of mmf by 30° . By considering more instants of time it will be seen that the movement of current phasor by a certain angle in the anticlockwise direction results in the rotation of the resultant mmf by the same angle in the clockwise direction.

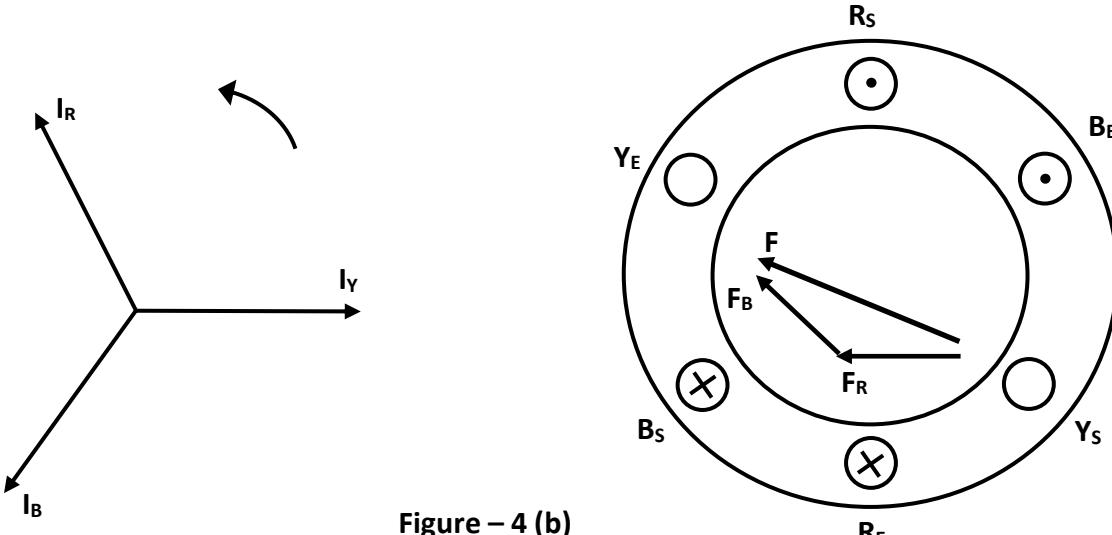


Figure – 4 (b)

Hence, it is clear that when three phase voltages are applied to three phase windings, the currents flowing through the windings produce a rotating magnetic field that is of constant amplitude and constant speed.

Construction: The two main parts of an induction motor are Stator and Rotor.

Stator: It consists of a steel frame which encloses a hollow, cylindrical core made up of thin laminations of silicon steel to reduce hysteresis and eddy current loss. A number of evenly spaced slots are provided on the inner periphery of the laminations. The insulated conductors are placed in the stator slots and are suitably connected to form a balanced three-phase star or delta-connected circuit. The three phase stator winding is wound for a definite number of poles as per the requirement of speed.

Rotor: The rotor is mounted on a shaft with a laminated core having slots on the outer periphery. Depending on the type of winding placed in the slots, we have two types of rotors - Squirrel cage rotor and Slip ring or wound rotor.

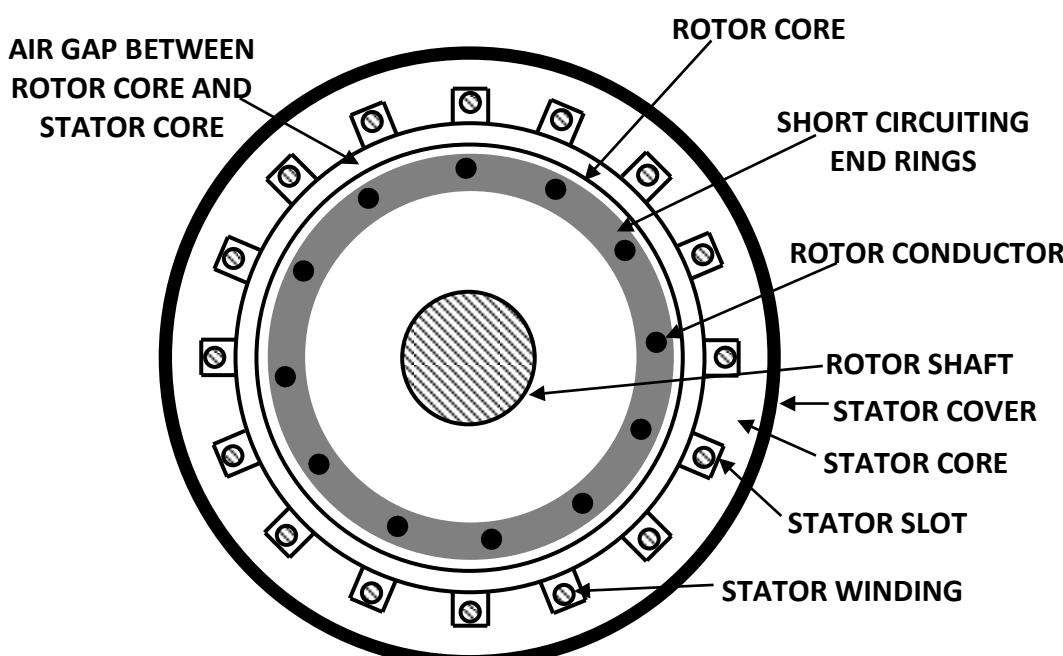


Fig. - 5 Constructional details of a squirrel cage induction motor

BASIC ELECTRICAL & ELECTRONICS ENGINEERING NOTES

Squirrel cage rotor: It consists of a laminated cylindrical core having parallel slots on its outer periphery. Copper or Aluminum bars are placed in each slot. All these bars are joined at either ends by copper or aluminum rings called end rings. This forms a permanently closed circuit winding. As the placing of bars and end rings resembles a squirrel cage it is called squirrel cage rotor. As the rotor circuit is a closed one external resistance cannot be added.

Wound rotor:

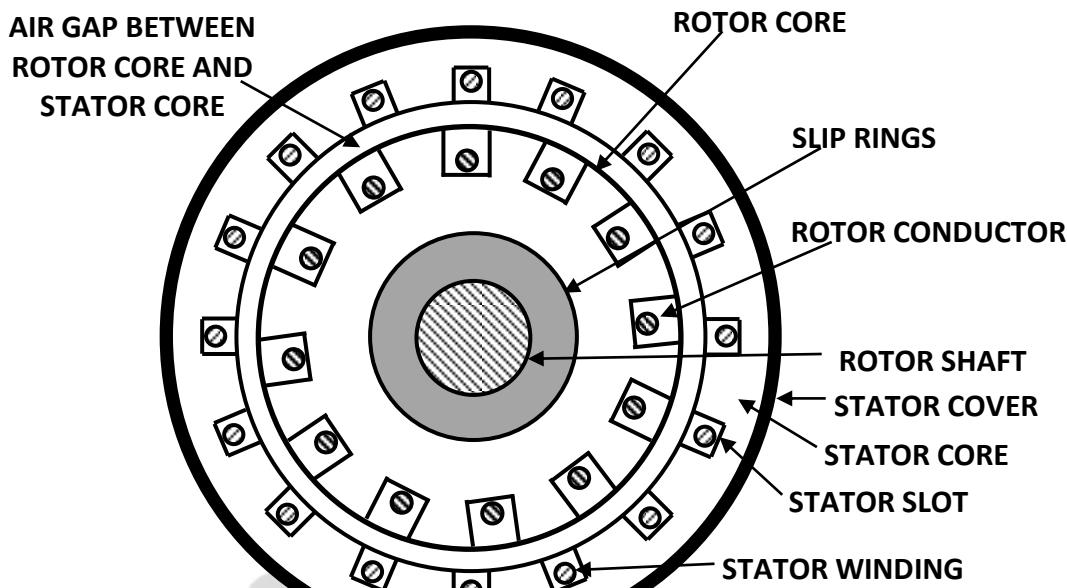


Fig. - 6 Constructional details of a slip ring induction motor

It consists of a laminated cylindrical core and carries a three phase winding similar to that on the stator. The rotor winding is uniformly distributed in the slots and is usually connected in star. The open ends of the rotor windings are brought out and connected to three insulated slip rings mounted on the rotor shaft with one brush resting on each slip ring. The three brushes are connected to a three phase star connected rheostat to vary the resistance of the rotor circuit. At starting the resistance is included to enhance the starting torque and gradually reduced as the motor picks up speed.

Working and Principle of operation: When three phase voltages are applied to the three phase stator winding, currents flow through the stator winding and a rotating magnetic field is set up. This rotating magnetic field rotates around the rotor at synchronous speed given by $N_s = 120f / P$. This rotating magnetic field passes through the air gap and cuts the rotor conductors that are initially stationary. Due to the relative speed between the rotating flux and the stationary rotor, emf's are induced in the rotor conductors. Since the rotor circuit is a closed one, currents start flowing in the rotor conductors. When current carrying conductors are placed in a magnetic field produced by the stator, the rotor conductors are acted upon by a mechanical force. The sum of the mechanical forces of all the conductors produces a torque that tends to move the rotor in the same direction as the rotating field.

BASIC ELECTRICAL & ELECTRONICS ENGINEERING NOTES

Concept of slip in three phase induction motor:

A three phase induction motor will function similarly as a transformer when viewed in a specific manner. The stator and rotor winding is always made for three phase and for a specific number of poles. The three phase induction motor at rest is similar to a transformer wherein the stator winding becomes the primary and rotor winding becomes the secondary.

The rotating magnetic field which rotates at synchronous speed is given by $N_S = \frac{120f_1}{P}$

Where N_S is Synchronous speed, which is the speed at which the stator magnetic field rotates

f_1 = Frequency of supply voltage impressed across the stator winding

P = Number of poles of induction motor

This rotating magnetic field cuts the rotor coils inducing emf and hence current as they form a closed circuit. The current carrying rotor coils in the midst of a rotating magnetic field experience mechanical forces which yield torque. As the rotor starts rotating, the rate at which the rotating field cuts the rotor coils decreases. This in turn reduces the induced emf per phase in the rotor coils. As the rotor picks up speed, it starts rotating at a speed lesser than the synchronous speed. As the current flowing in the rotor coils is produced only due to induction, it is necessary that there has to be a difference between the rotating magnetic field due to stator and the rotor. This difference in speed is called slip speed which is a basic requirement for the functioning of an induction motor.

We have, $\text{Slipspeed} = N_S - N \text{ rpm}$ Where N is the actual speed with which the rotor is rotating

The term slip describes the lagging behind of the rotor in comparison with the rotating magnetic field of the stator. The value of Slip for an induction motor lies between zero and unity.

The slip expressed on a per unit basis using synchronous speed as reference is -

$$\text{Fractional Slip, } S = \frac{N_S - N}{N_S} \quad \text{and} \quad \text{Percentage Slip, } \%S = \frac{N_S - N}{N_S} \times 100$$

When the induction motor is rotating, the emf induced in the rotor coils will be dependent on the slip S . The frequency of voltage or current induced in the rotor coils due to the relative speed or slip speed is given by $f_2 = Sf_1$

Where f_2 is the slip frequency or frequency of the induced emf in the rotor coils when the induction motor is rotating.

Torque-Slip characteristics of three phase induction motor:

When $S = 0, T = 0$, so the torque-slip characteristic starts from the origin. At normal speed, slip is small so the product of slip and rotor reactance per phase at stand still is negligible compared to R_2 .

As R_2 i.e. rotor resistance per phase is constant. We have $T \propto S$

The torque-slip curve is a straight line from zero slip to a value of slip that corresponds to full load.

As the slip increases beyond full load slip, the torque increases and becomes maximum at $S = \frac{R_2}{X_2}$

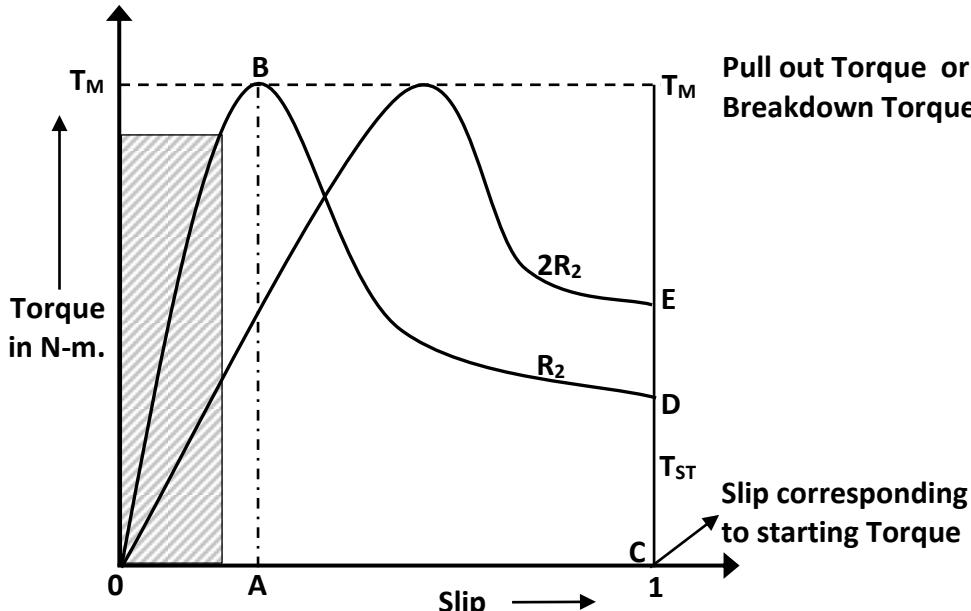


Fig. - 7 Torque - Slip Characteristics of Three phase induction motor

This maximum torque in an induction motor is called Pull out torque or Break down torque. When slip increases further due to load on induction motor the term $S^2 X_2^2$ increases very rapidly so that R_2^2 may be neglected as compared to $S^2 X_2^2$; $\therefore T \propto \frac{S}{S^2 X_2^2}$ or $T \propto \frac{1}{S}$ as X_2 is constant.

Now the torque is inversely proportional to slip so the torque-slip curve will resemble a rectangular hyperbola. It is observed that the addition of resistance to the rotor circuit does not change the value of maximum torque T_M but it only changes the value of slip at which T_M occurs.

Referring to the torque-slip characteristic, the hatched area shows the normal operating range for the induction motor. Within the region OB the operation of the motor will be stable and the range BD is the unstable region. CD represents the torque at $S = 1$; i.e. stand still condition or the starting torque.

The torque-slip characteristics for a slip range from $S = 0$ to $S = 1$ for some values of rotor resistance are also shown in fig. - 7.

Applications: Squirrel cage induction motors are simple and rugged in construction, cheap and require less maintenance. They are preferred for many industrial applications. They are used in - Lathes, Drilling machines, Industrial and Agricultural pumps, Compressors and Industrial drives.

Slip ring induction motors when compared to squirrel cage induction motors have high starting torque, smooth acceleration under heavy loads, adjustable speed and good running characteristics. They are used in - Lifts, Cranes and Conveyors.

SYNCHRONOUS MACHINES

Synchronous machines are three phase electrical machines which rotate at synchronous speed. A synchronous machine can work as a motor as well as a generator. There is not much constructional difference between a synchronous motor and a synchronous generator. The stator core will be identical but the rotors shall be different.

BASIC ELECTRICAL & ELECTRONICS ENGINEERING NOTES

Construction: The main parts of a synchronous machine are - Stator frame, Stator core and Rotor. The stator frame usually holds the stampings of the stator core and windings in position. The frame is cast for small machines and for large machines the frame is fabricated from mild steel plates. Holes are provided in the stator frame for ventilation as the stator core has radial ventilating spaces provided in the stampings.

Stator core: It is made up of laminations of special magnetic iron or steel alloy. The core is laminated to minimize eddy current loss. The lamination may be a single stamping or made up of segments. The laminations are insulated from one another and have radial spaces between them for allowing the cooling air to pass through. The slots may be open or semi closed in which the windings of the stator are placed.

SYNCHRONOUS GENERATOR

Rotor for generator: Rotor used in synchronous machines which work as a generator are of two types - Salient pole and Non - Salient pole.

Salient pole or Projecting pole type: The rotor has laminated projecting poles shaped in a specific manner and fixed to the cast iron rotor wheel. The projecting pole makes provision for placing the field coil for DC excitation. The individual field windings are connected in series in such a way that when energized, adjacent poles become north and south. The DC supply to the field coils is fed through brushes sliding over slip rings fixed to the shaft of the rotor. These rotors have large diameters and short axial lengths. This type of rotor is employed for slow and medium speed machines mostly driven by diesel engines.

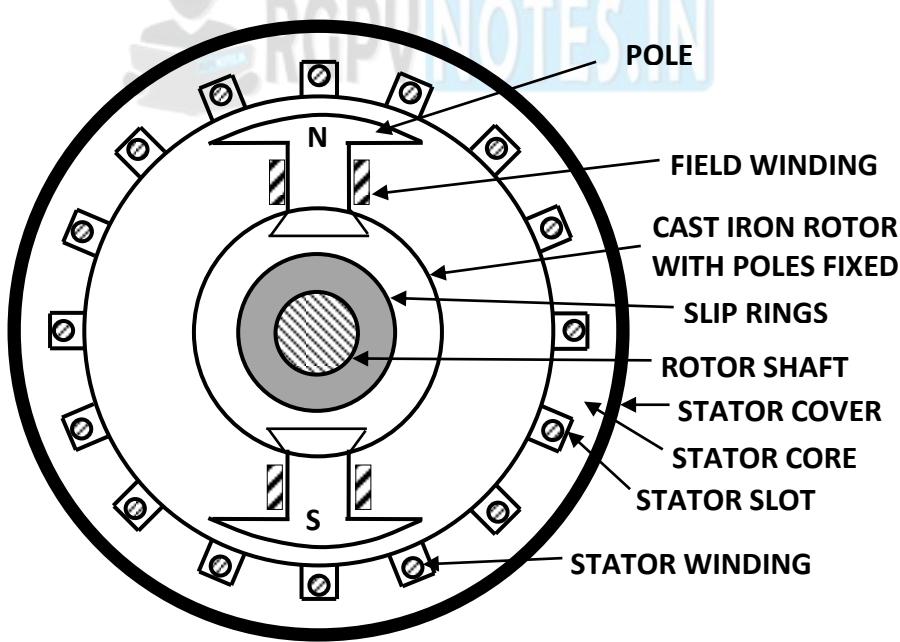


Fig. - 8 Constructional details of a salient pole synchronous machine

Non-Salient or smooth cylindrical type: The rotor is made of smooth solid forged steel radial cylinder having a number of slots along the outer periphery. The field winding is placed in these slots and connected in series. The un-slotted part of the rotor forms the pole faces. These rotors have smaller diameters and large axial length. The number of poles is restricted to 2 or 4 poles. The windings are placed in the slots around the pole faces in such a way that, the flux density is maximum on the polar central line and decreases on

BASIC ELECTRICAL & ELECTRONICS ENGINEERING NOTES

either side. DC supply is given to the field winding through slip rings and brushes so that alternate north and south poles are formed. These rotors are used for high-speed machines driven by turbines.

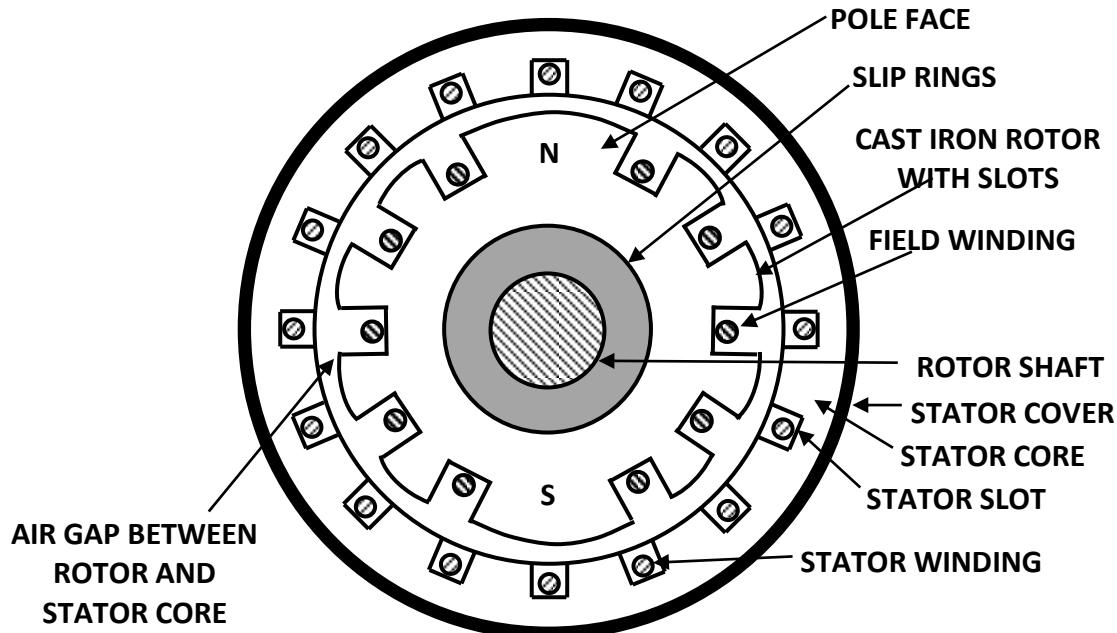


Fig. - 9 Constructional details of a non-salient pole synchronous machine

Working and Principle of operation as a generator: A synchronous generator operates on the fundamental principle of electromagnetic induction. It has a field winding on the rotating part and is fed with DC supply through two slip rings. The field windings develop alternate N and S poles on the rotor. The armature winding which is made for three phase is placed in the slots of the stator which is stationary. When the rotor is rotated by a prime mover, the magnetic flux of rotor poles cut the armature conductors placed in the slots of the stator. Consequently emf is induced in the armature conductors due to Electromagnetic Induction. The induced emf is alternating since N and S poles of rotor pass under the armature conductors. The direction of induced emf can be found by Fleming's right hand rule and the frequency is given by $f = PN/120$. The magnitude of the induced emf in each phase of the armature winding is the same but they differ in phase by 120° .

Emf equation:

Let Z = Number of conductors in series per phase, Φ = Flux per pole in weber

P = Number of rotor poles, N = Rotor speed in rpm

In one revolution (i.e. $60/N$ sec) each stator conductor is cut by $P\Phi$ weber

$$\text{i.e. } d\Phi = P\Phi \text{ and } dt = 60/N$$

$$\text{Average emf induced in one stator conductor} = \frac{d\Phi}{dt}; \quad \therefore \frac{d\Phi}{dt} = \frac{P\Phi}{60/N} = \frac{P\Phi N}{60}$$

As there are Z conductors in series per phase

$$\text{Average emf per phase} = \frac{P\Phi N}{60} \times Z; \text{ we have } f = \frac{PN}{120} \quad \text{or} \quad N = \frac{120f}{P}$$

BASIC ELECTRICAL & ELECTRONICS ENGINEERING NOTES

$$\text{Average emf per phase} = \frac{P\Phi Z}{60} \times \frac{120f}{P} = 2f\Phi Z \text{volts}$$

rms value of emf per phase = Average emf per phase × Form factor = $2f\Phi Z \times 1.11$

$$E_{rms}/\text{phase} = 2.22f\Phi Z \quad \text{or} \quad E_{PH} = 2.22K_p K_d f\Phi Z$$

where K_p - Pitch factor and K_d - Distribution factor.

$$\text{Finally we have } E_L = \sqrt{3}E_{PH}; \quad \therefore E_L = \sqrt{3} \times 2.22K_p K_d f\Phi Z$$

Applications: Synchronous generators are used in all power plants for generation of Electricity.

SYNCHRONOUS MOTOR

Rotor for synchronous motor: There are two types of rotors - With amortisseur winding on rotor pole faces and without amortisseur winding. The rotor used for the motor has laminated projecting poles shaped in a specific manner fixed to the cast iron rotor wheel. The projecting poles make provision for placing the field coil for DC excitation. The pole faces have slots for placing the amortisseur winding for starting. The amortisseur winding consist of copper or bronze bars embedded in the slots and short circuited at both ends by conducting rings, similar to the squirrel cage winding to develop starting torque by induction motor action. This winding also serves to damp out oscillations of speed during normal operation.

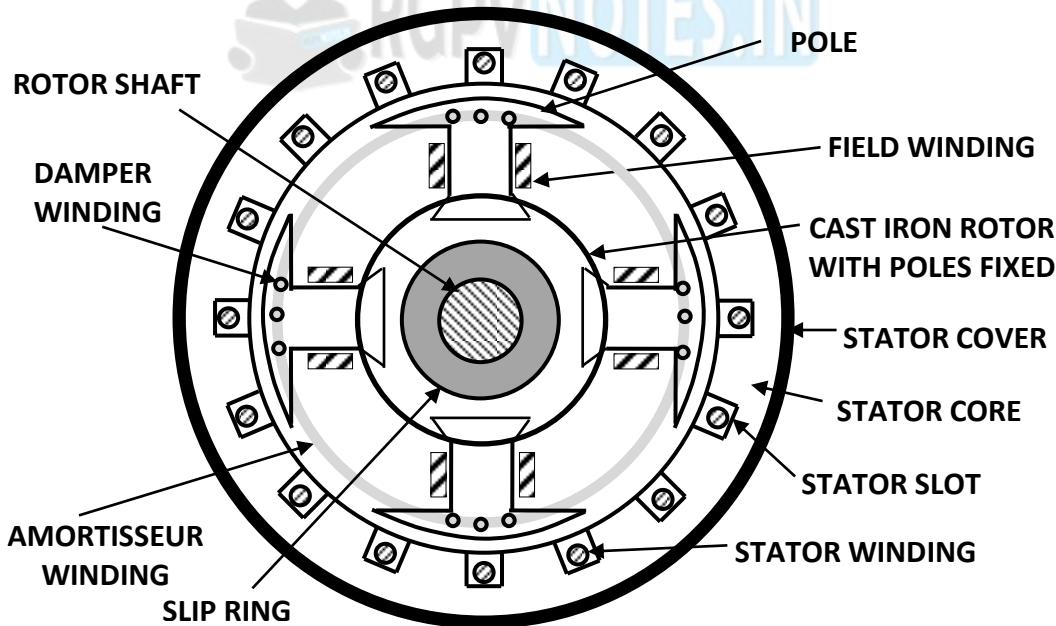


Fig. - 10 Constructional details of a synchronous machine with ammortisseur winding

Working and Principle of operation as a motor: The stator of a synchronous motor when fed by a three phase AC supply, produces a magnetic flux of constant magnitude but rotating at synchronous speed. The rotor pole faces have the amortisseur winding which helps in developing the starting torque by induction motor action. As the motor approaches the synchronous speed, the rotor field winding is excited by a DC source which creates alternate N and S poles on the rotor. The poles of the rotor will face the poles of

BASIC ELECTRICAL & ELECTRONICS ENGINEERING NOTES

opposite polarity on the stator and a strong magnetic attraction is setup between them. The rotor poles lock in with the poles of rotating flux. Consequently the rotor rotates at the same speed as the stator field i.e. at synchronous speed. When the rotor rotates at synchronous speed, i.e. the same speed as that of rotating magnetic field, there will be no emf induced in the amortisseur winding. Hence due to the magnetic locking the synchronous motor can only run at synchronous speed.

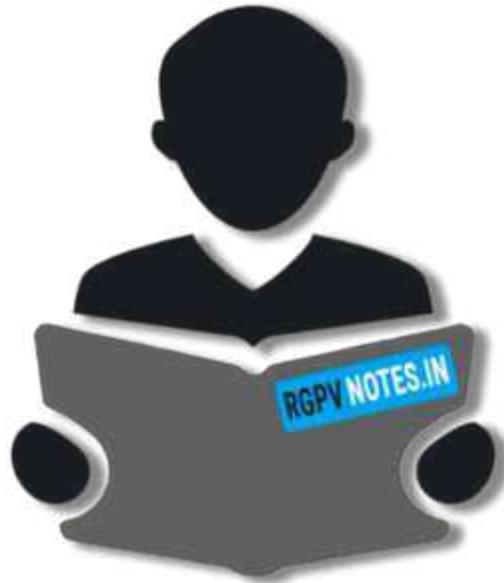
Applications: Synchronous motors with speeds below 500 rpm are used in Electroplating, Reciprocating compressors, Centrifugal pumps, Rolling mills and Paper mills.

Synchronous motors with speeds above 500 rpm are used in Fans, Blowers and Frequency changers. Over excited Synchronous motors on no load draw leading power factor current which can be used in Electrical sub-stations to improve power factor of the power system.

Types of losses occurring in Electrical machines:

The losses occurring in electrical machines can be listed as:

1. Constant losses
2. Losses occurring in the machine which are proportional to the current drawn by the machine
3. Losses in machines which are proportional to the square of the current drawn by the machine
 - ❖ Losses can be classified as Constant losses and Variable losses
 - ❖ Constant losses comprise of Core losses and Mechanical losses
 - ❖ Core losses are made up of Hysteresis and Eddy current loss
 - ❖ Mechanical losses comprise of Windage and Friction loss
 - ❖ Variable losses comprise of Copper losses and Stray load losses
 - ❖ Copper losses comprise of Stator copper loss and Rotor copper loss
 - ❖ Stray load losses can occur in the core as well as in the winding



RGPVNOTES.IN

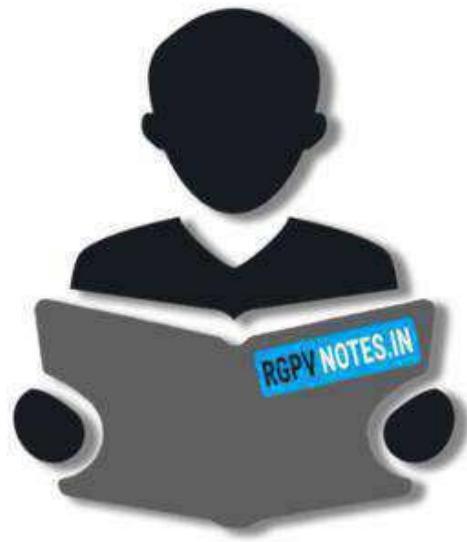
We hope you find these notes useful.

You can get previous year question papers at
<https://qp.rgpvnotes.in> .

If you have any queries or you want to submit your study notes please write us at
rgpvnotes.in@gmail.com



LIKE & FOLLOW US ON FACEBOOK
facebook.com/rgpvnotes.in



RGPVNOTES.IN

Program : **B.Tech**

Subject Name: **Engineering Chemistry**

Subject Code: **BT-101**

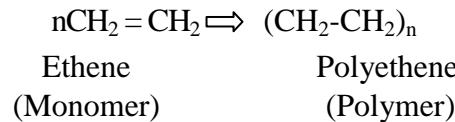
Semester: **1st**



LIKE & FOLLOW US ON FACEBOOK
facebook.com/rgpvnotes.in

POLYMERS

The word 'polymer' is derived from the classical Greek words poly meaning 'many' and 'meres' meaning 'parts'. Simply stated, a polymer is a molecular compound with high molecular mass, ranging into thousands and millions of gram. It is composed of a large number of repeating units of identical structure called monomers. For example, polythene is a polymer formed by linking together of a large number of ethene molecules.



Certain polymers, such as proteins, cellulose, and silk are found in nature and hence are called natural polymers, whereas, a number of polymers are produced by synthetic routes and are called synthetic polymers. In some cases, naturally occurring polymers can also be prepared synthetically. For e.g. rubber, which is also known as polyisoprene?

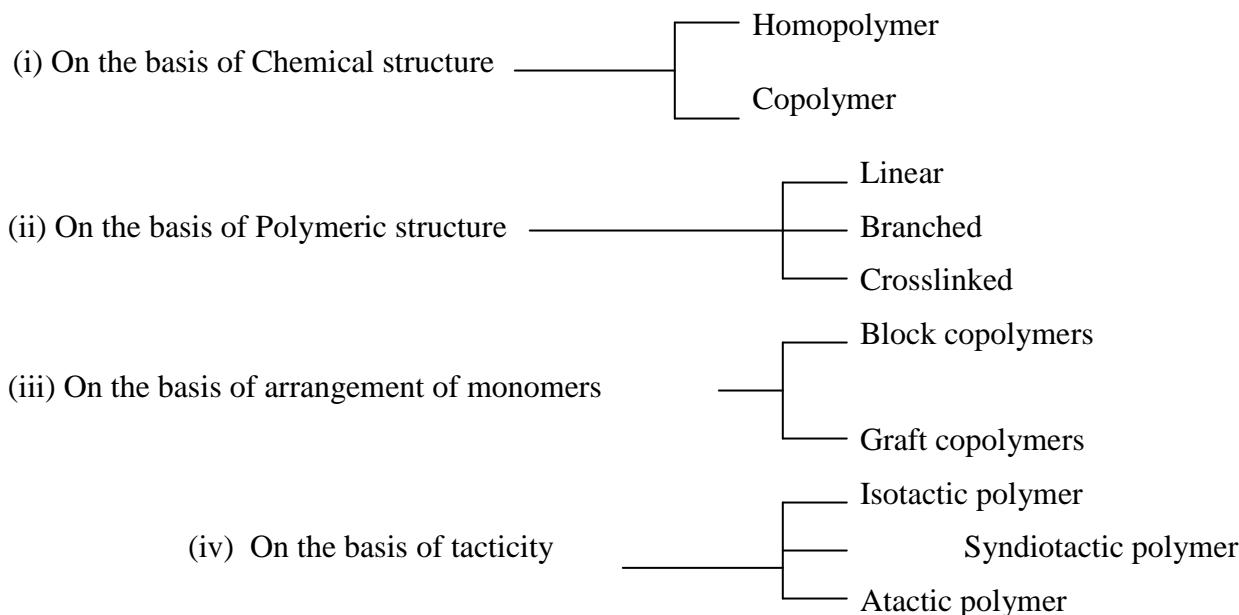
Both natural and synthetic polymers impinge on every aspect of our existence. The idea is often expressed that we live in "a plastic age". The range of polymeric materials is so wide that they have found applications in practically every branch of industry. About 80% of today's chemists, including biochemists, work with polymers.

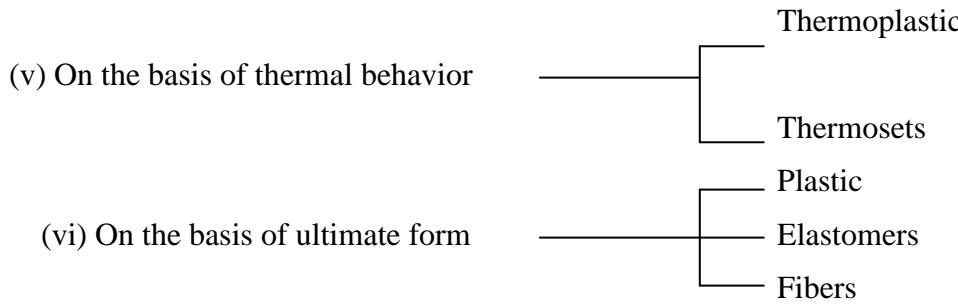
The number of repeating units in a chain formed in a polymer, is known as the "degree of polymerization". Polymers with high degree of polymerization are termed as "high polymers" and those with low degree of polymerization are called oilgopolymers.

CLASSIFICATION OF POLYMERS

Synthetic polymers have been classified into various groups, in various manners.

CLASSIFICATION OF POLYMERS

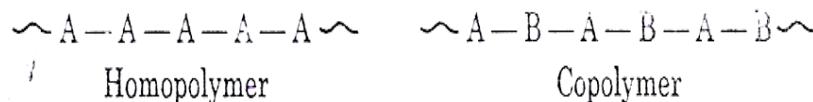




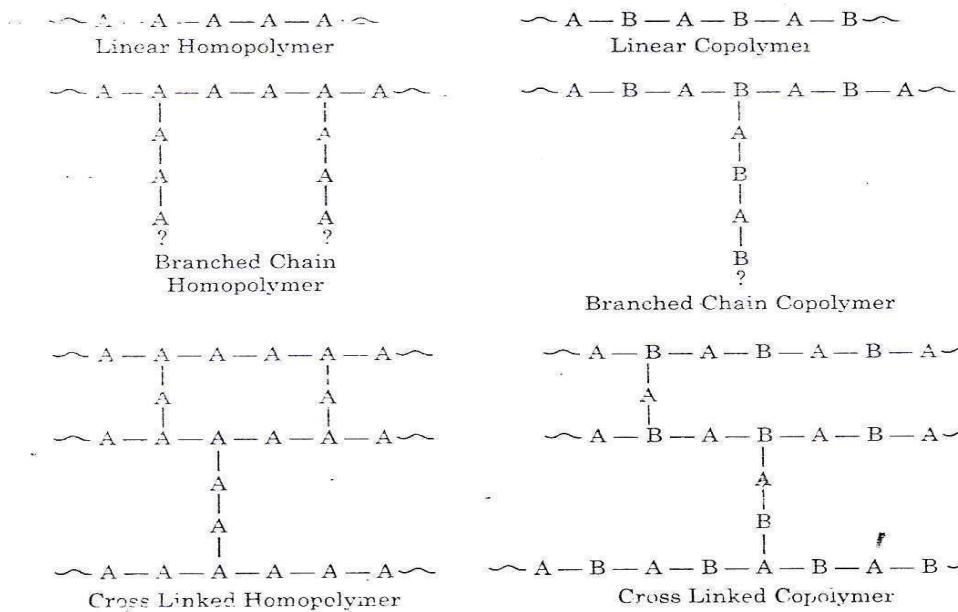
(i) On the basis of the chemical structure of their backbone, polymers are of two types:

Homopolymer: If a polymer consists of identical monomer, the polymer is termed as homopolymer.

Copolymer: If the polymer is a mixture of more than one type of monomer it is termed as copolymer.



(ii) On the basis of their polymeric structures, the polymers can be classified as linear, branched or Crosslinked.



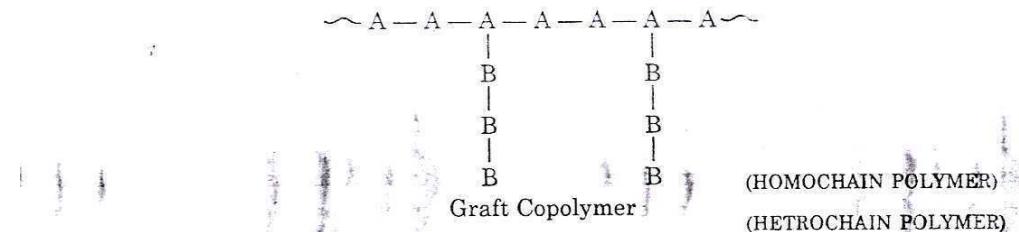
(iii) Depending upon the arrangement of the monomers in the case of copolymers, the polymers are divided into block copolymers and graft copolymers.

Block Copolymers: Linear polymers in which the identical monomeric units occur in relatively long sequences are called block co-polymers.



Graft Copolymers: Graft copolymers are branched copolymers in which the backbone is formed from one type of monomer and branches are formed of the other, i.e., the monomer

segments on the branches and backbone are not the same.

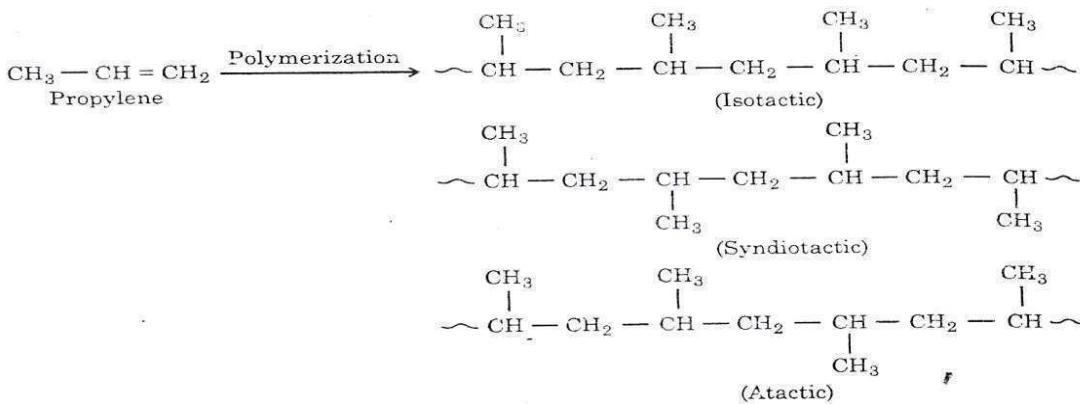


- (iv) Depending upon the orientation of monomer units in a polymer molecule: With respect to the main chain the polymers are classified as:

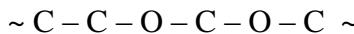
Isotactic polymer: If the side groups of the monomers lie on the same side of the chain, it is called an “isotactic” polymer.

Syndiotactic polymer: If the side groups are arranged in an alternate fashion, it is called syndiotactic polymer.

Atactic polymer: If the side groups are arranged in irregular fashion on randomly around the main chain, it is called atactic polymer.



- (v) If the main chain is composed of the atoms of the same specie, the polymer is called homochain and if the main chain is made up of different atoms then they are called heterochain polymers.



- (vi) According to their behavior on heating, the polymers are divided into two groups.

Thermoplastics: Those polymers, which soften on heating and become plastic so that they can be converted to any shape by moulding, are known as thermoplastics.

Thermosets: Those polymers which change irreversibly into hard and rigid materials on heating and cannot be reshaped, once they are set are called thermosets. They form infusible and insoluble mass on heating.

Plastics: All the synthetic polymers (other than elastomers) are usually referred to as plastics. The polymers, which are shaped into hard and tough utility articles by the application of heat and pressure, are called plastics. At some stage of manufacture, these are in plastic condition (Plasticity is the property by virtue of which a material undergoes permanent deformation under stress).

Elastomers: The polymers which are capable of being stretched rapidly at least 150 percent of their original length without breaking and return to their original shape on release of stress is known as elastomers. Synthetic rubbers constitute important of this group.

FUNCTIONALITY

The functionality of a molecule is the number of reactive sites it has. For a substance to act as a monomer, it must have two reactive sites. In other words, they should have functionality of at least two. A compound assumes functionality because of the presence of reactive functional groups present in a compound defines its functionality. For example –

Compound	Chemical Formula	Functionality
Acetic Acid	CH ₃ COOH	1
Molonic acid	HOOC CH ₂ COOH	2
Ethyl alcohol	C ₂ H ₅ OH	1
Ethylene glycol	HOCH ₂ CH ₂ OH	2
Lactic acid	CH ₃ CH(OH) COOH	2
Tartaric acid	HOOC(CHOH) ₂ COOH	4

There are some other compounds in which the presence of easily replaceable hydrogen atoms imparts functionality. For example, phenol, it has got –OH groups as a functional group but can also undergo substitution reaction at these three sites replacing its three hydrogen atoms. For example, in reaction (i), phenol exhibits mono functionality and in reaction (ii), it exhibits a functionality of three.

Depending on the functionality of the monomers used, we get linear, branched or three dimensional cross-linked polymers.

TYPES OF POLYMERIZATION

There are two types of polymerization processes:

1. Addition Polymerization,
2. Condensation Polymerization.

Addition Polymerization

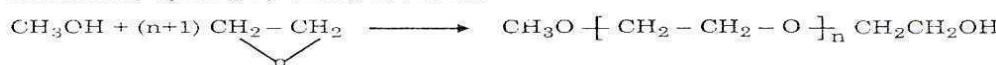
Addition polymerization is characterized by self-addition of the monomer molecules to each other, without the loss of any material. The product is an exact multiple of the original monomeric molecule.

Examples:

- (i) Addition by breaking up of multiple bonds:



- (ii) Addition by opening up of ring structures:



Vinyl compound ($\text{CH}_2 = \text{CHX}$), allyl compounds ($\text{CH}_2 = \text{CHCH}_2 \text{X}$), olefins ($\text{CH}_2 = \text{CHR}$) and dienes ($\text{CH}_2 = \text{CR}-\text{CH}=\text{CH}_2$) typically undergo addition polymerization. Since a majority of these fall under the 'vinyl' category, addition polymerization is also customarily termed as vinyl polymerization.

MECHANISM OF ADDITION POLYMERIZATION

Addition polymerization is a typical chain reaction and consists of three important steps:

(i) Initiation of the formation of active centers.



(ii) Chain propagation



(iii) Chain termination or removal of active centre.



The process of addition polymerization can be brought about by three methods:

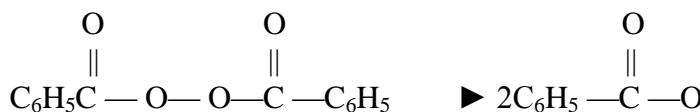
1. Free radical mechanism
2. Ionic mechanism (Anionic and Cationic)
3. Coordination mechanism

The three types of mechanisms for addition polymerization, depending upon the active centres formed, are discussed below.

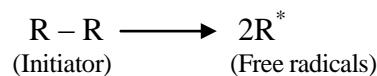
1. **Free Radical Addition Polymerization:** The free radical addition polymerization involves the formation of free radicals in the initiation step, which can be produced by decomposition of compounds called initiators. The free radical polymerization has three principal steps:

(i) **Initiation:** Initiation in a free radical polymerization consists of two steps. The first is the 'generation of the free radical species, followed by the second step which involves the addition of a single monomer molecule to the free radical.

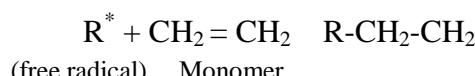
The generation of free radical takes place as a result of homolytic dissociation of an initiator, which may take place in the presence of heat energy, light energy or catalysts. A number of low molecular weight compounds, comprising mainly of azo compounds, peroxides, hydroperoxides, peracids and peresters are useful as initiators. An important example of a free radical initiator is benzoyl peroxide which dissociates as:



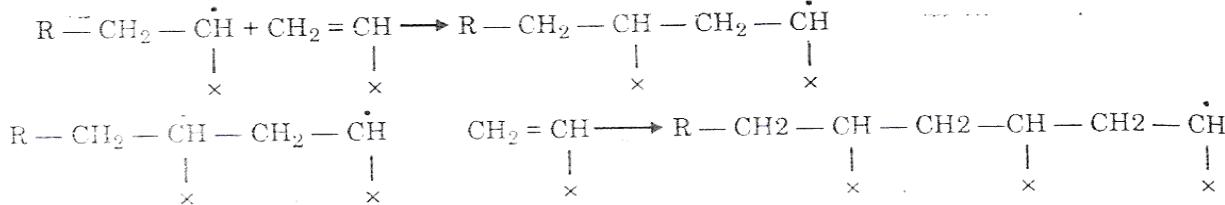
In general, the dissociation of the initiator (R-R) to form free radical specie can be represented as



In the second step of initiation, the free radical so formed gets attached to the monomer molecule (M) to form the chain initiating species



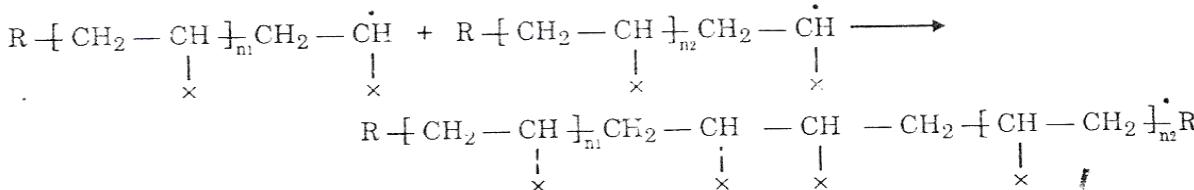
(ii) **Propagation:** In the next step, additional monomer units are added to the initiated monomer species, as



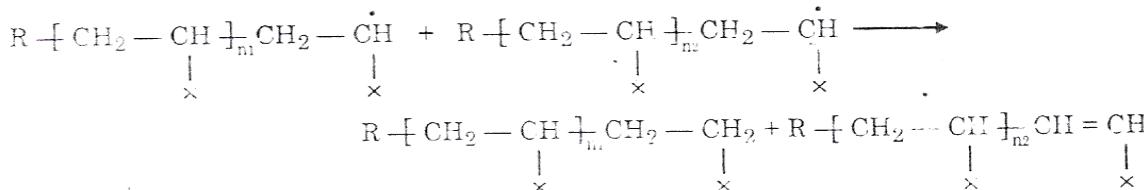
The propagation step involves a continuing attack on fresh monomer molecules which, in turn, keep successively adding to the growing chain one after another.

- (iii) **Termination:** Propagation will continue until some termination process takes place. The propagating polymer chain is terminated by any of the following reactions.

(a) *By combination:* Since the decomposition of the initiator produces many free radicals at the same time, each one of them can initiate and propagate the growth of a number of chains simultaneously. The two growing chains may come close and collide with each other and get deactivated.

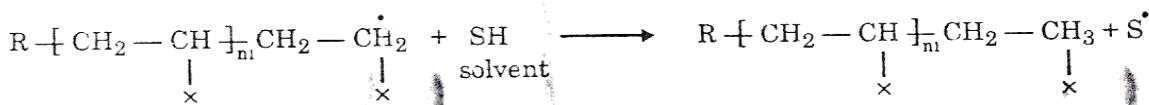


(b) *By disproportionation:* In some cases, one hydrogen atom from one growing chain is abstracted by the other growing chain and results in the formation of two polymer molecules, one saturated and other unsaturated. In this case the termination process results in the formation of two polymer molecules of shorter chain length as against a single molecule of a longer chain length obtained by first method.



Thus the products so formed do not contain any reactive sites and may be termed as dead polymers as these cannot grow further.

- (c) In addition to the above two methods of chain termination, chain transfer by hydrogen abstraction from solvent molecule, may also terminate the growing chain.



In this reaction the original growing chain is terminated whereas the new chain gets initiated by the solvent free radical, hence this is called termination by chain transfer.

Certain chemical compounds, which are capable of inhibiting or completely stopping the chain growth process, are called radical inhibitors. Phenolic compounds are widely used as inhibitors, because phenolic hydrogen can be easily abstracted by the polymer chain. The phenolic free radical so formed is highly stabilized and no further reaction is possible. Hydroquinone is often added as inhibitor.

Polythene, polypropylene, styrene etc. are some of the polymers that can prepare by tree

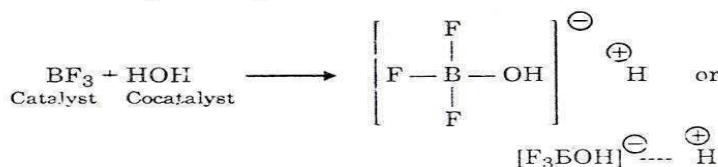
radical polymerization. Branching and cross linking during the free-radical addition polymerization.

2. Ionic Addition Polymerization: Ionic polymerization follows the same basic steps as free-radical addition polymerization i.e. initiation, propagation and termination, in this process, the initiation step involves the formation of carbonium ion or the carbanion as the active site and hence the polymerization can be termed as cationic or anionic polymerization respectively.

(i) Cationic Polymerization

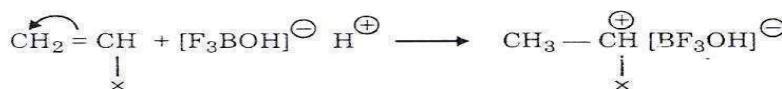
Cationic polymerization takes place in the presence of catalyst that is regenerated at the end of polymerization (in contrast to free radical and anionic polymerization where the catalyst added becomes incorporated into the terminated Polymer chain). Polymerization of vinyl monomers with an electron-donating group (such as $-OC_2H_5$ - OCH_3 - C_6H_5 etc.) may take place by a cationic mechanism which involves the following steps.

Initiation: This step involves the formation of a carbonium ion by the attack of a proton on the monomer. Any strong Lewis acid like BF_3 , $AICl_3$ etc can be used as a catalyst. In this case a cocatalyst like water is required to provide the proton source.

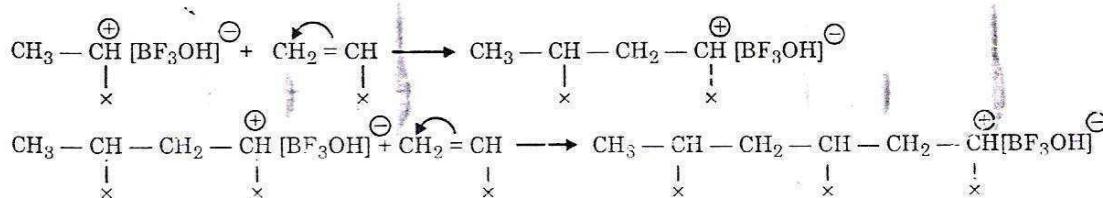


H^+ is the initiator and $[F_3BOH]^\ominus$ is the counter ion.

The H^+ now attacks the n electron of the monomer and forms the carbonium ion. The counter ion $[F_3BOH]^\ominus$ anion faces the growing chain end.



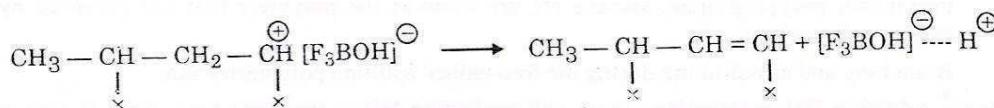
Propagation: The propagation reaction involves the addition of monomer units and simultaneous transfer of the charge to the newly added monomer unit.



More and more units are added up and the chain keeps on growing. The electron pair of the adding monomer units are pulled in a direction opposite to the growth of the chain and the positive charge keeps on moving in the direction of the chain growth. The counter ion $[BF_3OH]^\ominus$ moves along the positive charge all the time.

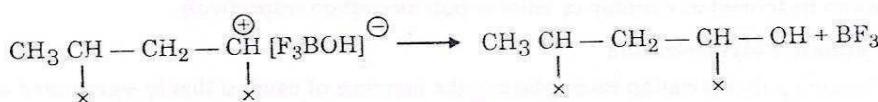
Termination: Termination takes place when the growing polymer chain having C^+ collides with an anion, which may be added deliberately or with the counter ion. The possible reactions may be:

(a) Donation of a portion to the counter by the growing polymer chain. This results in the formation of a double bond at the end of growing polymer molecule.



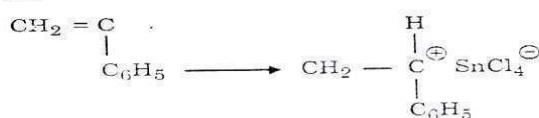
Thus the initiator or catalyst is regenerated.

(b) Formation of covalent bond between the carbonium ion and the counterion (coupling):

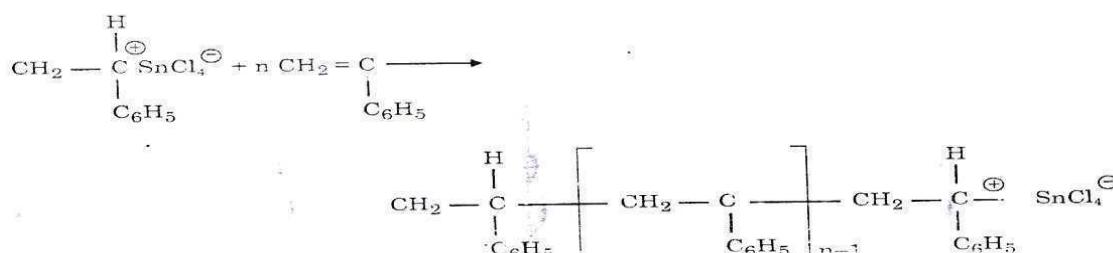


In this case also the regeneration of the initiator takes place.

Examples of the monomers that can undergo cationic polymerization are isobutylene, styrene methyl styrene and many vinyl others. Cationic

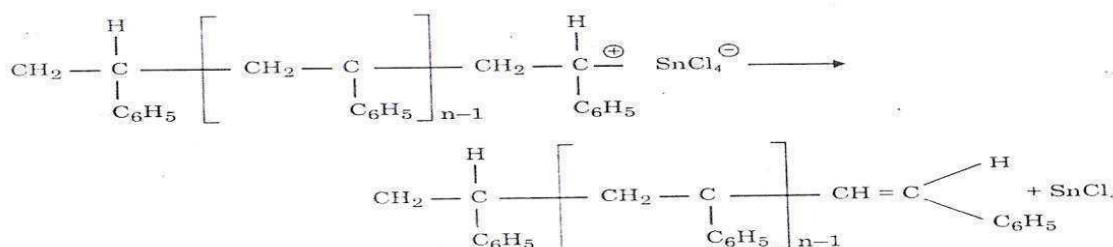


Propagation: Carbonium ion attacks the another molecule of styrene.



More and more styrene molecules are added and the charge is simultaneously transferred to the newly added monomer unit.

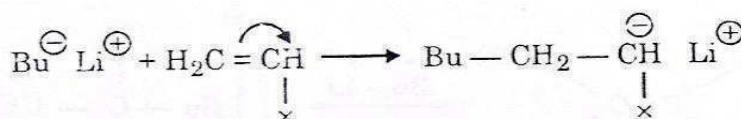
Termination: Termination takes places by donation of a proton to the counter ion by growing polymer chain forming a double bond



(ii) Anionic Polymerization

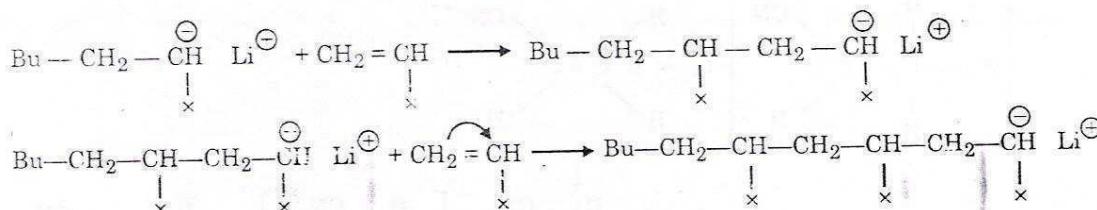
Anionic polymerization involves the formation of the carbanion in the initiation step. Polymerization of vinyl monomers with an electron withdrawing group can proceed by an anionic mechanism, involving the following steps:

Initiation: The initiators or the catalysts used in anionic addition polymerization are strong bases such as organo alkali compounds (alkyl or aryl derivatives of alkali metals) like n-butyl lithium, ethyl sodium etc. Alkali metal amides, hydroxides etc. can also be used.



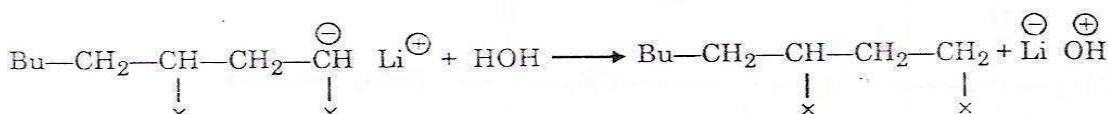
The electron pair in this case is pushed to the end of the molecule forming the carbanion. Li^\oplus is the counter ion.

Propagation: The carbanion formed, propagates the chain growth by attacking the second monomer **PAI** unit, pushing its electron pair further away to the end and forming a sigma bond with the new monomer unit.



In the case of anionic polymerization the movement of the **PAI** electrons is towards the direction of the chain growth (in cationic polymerization, the movement of the **PAI** electrons is opposite to the chain growth).

Termination: The termination in anionic polymerization is not a spontaneous process. If the starting reagents are pure and no impurities are present, propagation can proceed indefinitely or until all the monomer is consumed. Termination is generally accomplished by the transfer of negative charge to a species which is not directly involved in the reaction. For example CO_2 , methanol, water etc.

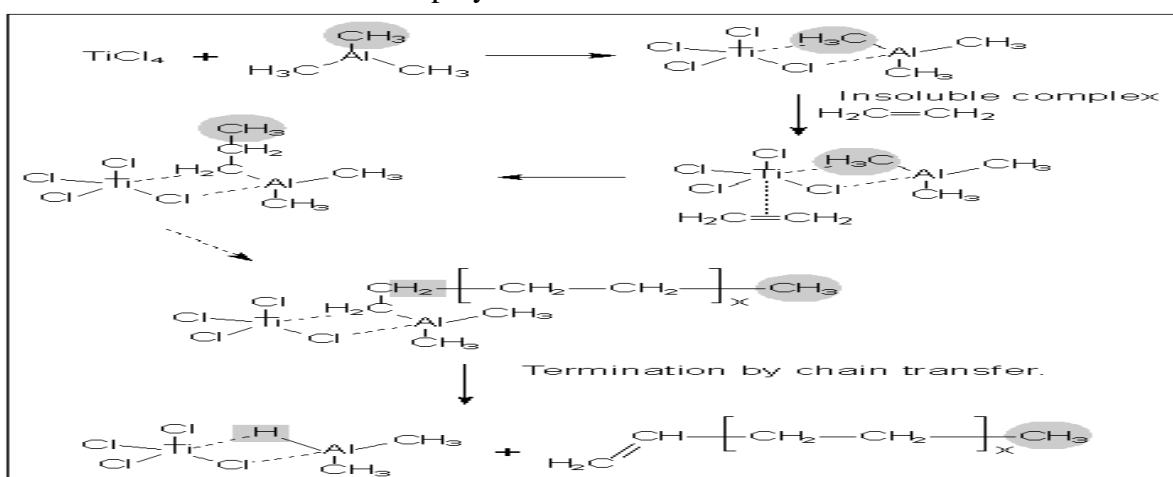


3. Coordination Polymerization: Polymerization reactions, taking place in the presence of organometallic compounds as catalysts, are termed as coordination polymerizations.

Zeigler (1953) and Natta (1955) discovered that in the presence of a catalyst which is a combination of a transition metal halide (like TiCl_4 or TiCl_3) with organo-metallic compounds (like triethyl aluminium), stereoregular polymers can be prepared. The combination is often referred to as, ZeigherNatta catalyst. The commonly used Zeigler-Natta is trhethylaluminium in combination with titanium trichloride.

In the formation of the monomer catalyst complex, a co-ordinate bond is involved between a carbon atom of a monomer and metal atom of the catalyst that is why, this type of polymerization is known as coordination polymerization. The co-ordinate metal-carbon bond formed in the monomer-catalyst complex acts as an active center from where propagation starts.

Mechanism of coordination polymerization can be illustrated as follows:



Coordination polymerization is used to prepare sterospecific polypropylene; polydiene etc. polypropylene of nearly 90% isotacticity can be prepared. The exact mechanism of coordination polymerization is still unclear.

Condensation Polymerization

Condensation polymerization is brought about by monomers containing two or more reactive functional groups with the elimination of small molecules like O, HCl etc. For example, It is seen from the above reaction that the polymer formed still contains both the reactive functional groups at its chain ends and hence, is 'active' and not 'dead'. In the above example, the different type of reactive functional groups, viz. $-NH_2$ and $-COOH$ are present on two different monomers. In some cases, both the reactive groups may be present on the same monomer and hence lead to self-condensation reaction. For example, polycondensation of amino acids.

Some polycondensation reactions leading to different types of polymers are:

- (i) Polycondensation of dihydric alcohols and dicarboxylic acids leads to the formation of **polyesters**.
- (ii) Polycondensation of dicarboxylic acids and diamines to give **polyamides**.
- (iii) Polycondensation of dihydric alcohols and methyl esters of dicarboxylic acids to give **polyesters**.
- (iv) Self condensation of $-OH$ yields **polyesters**.
- (v) Self condensation of amides to form **polyamides**.
- (vi) Polycondensation of bisphenol A and phosgene to form **polycarbonate**.

COPOLYMERIZATION

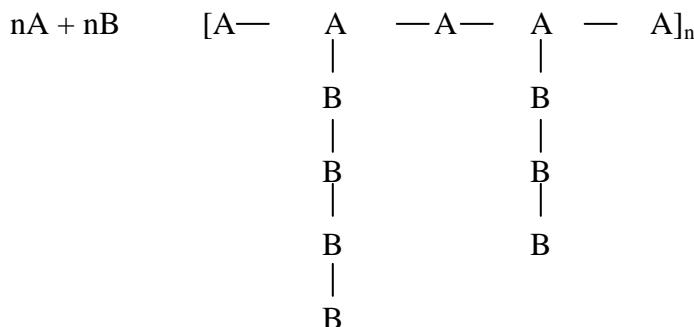
Commonly available nature and synthetic polymer suffer from many disadvantages and therefore cannot be used for special purposes. However, it is possible to modify a number of natural and commercially available synthetic polymers to give products which can be utilized for performing desired functions.

Copolymerization is the most general and powerful method of affecting modification in polymers and is widely used in the production of commercial polymers. It provides a technique to combine the properties of two or more different polymers into a single polymer. Copolymerization can be defined as any process whereby two or more monomers are copolymerized to form a single polymer. The resulting high molecular weight compound is known as copolymer. For example, butadiene and styrene, copolymerize to form butadiene-styrene rubber.

Depending upon the sequence in which different polymeric units are arranged, copolymers are of following types:

- (a) Random copolymers
 - (b) Alternating copolymers
 - (c) Block copolymers
 - (d) Graft copolymers
- (a) **Random copolymers:** In random copolymers the two monomers A and B are distributed randomly along the polymer chain.
- $$nA + nB \quad [A - A - B - B - A - B - A - A]_n$$
- (b) **Alternating copolymers:** In block copolymers, a sequence or block of one monomer is followed by the block of the other monomer.
- $$nA + nB \quad [A - B - A - B - A - B - A]_n$$
- (c) **Block copolymers:** In block copolymers, a sequence or block of one monomer is followed by the block of the other monomer.
- $$nA + nB \quad \begin{array}{ccccccccc} & [A - A - & A - B - & B - B - & B - A - & A - A - & A] \\ & \text{Block} & \text{Block} & & & & & \end{array}$$
- (d) **Graft copolymer:** Graft copolymers are branched polymers and contain main chain

or backbone consisting of one monomer and branches formed from the other monomer:



PLASTICS

As was discussed earlier, plastics area group of artificially synthesized materials, which can be moulded into any desired form, when subjected to heat and pressure in the presence of a catalyst.

The resins are the basic binding materials, which form the major part of the plastics and determine the type of treatment needed in the moulding operations.

From engineering point of view there are two types of resins –

- (i) Thermoplastic resins.
- (ii) Thermosetting resins.

(i) Thermoplastics: Thermoplastics or resins soften on heating and become plastic so that they can be converted to any desired shape by moulding. On cooling, they become hard and rigid but their hardness is a temporary property subject to change with increase or decrease in temperature. In thermoplastic resins, the chemical structure or the molecular weight are not changed during the heating or moulding operations, because the changes involved are of physical nature. Only the secondary bonds between the individual molecular chains are broken on heating which results in their softening. On cooling, these secondary bonds are again formed and hence they become hard again. Some common examples of thermoplastics are polyethene, polypropylene, polystyrene, polyvinyl chloride, teflon, polyacrylonitrile etc.

Plasticizer: Sometimes, the substances called plasticizers are added to improve plasticity and flexibility of the plastics. A variety of organic materials are used as plasticizers including non-drying vegetable oils, tributyl phosphate, triphenyl phosphate, triacetin camphor, esters of oleic or stearic acid.

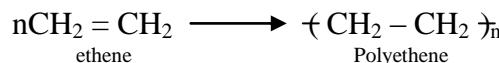
(ii) Thermosetting resins: Thermosetting resins are those, which upon heating, change irreversibly into hard and rigid materials on heating. After cooling, if the set article is again heated, it will not soften again. The thermosetting resins during moulding acquire three-dimensional cross-linked structures with predominantly covalent bonds, which retain their strength even on heating. Thus, the three dimensional network structure joined by strong covalent bonds make the thermosetting resins harder, stronger and more brittle than thermoplastic resins, if heating of a thermoset resin is prolonged, these undergo change in chemical composition to give a hard, infusible and insoluble mass. Some common examples include phenol formaldehyde, malamine formaldehyde, alkyds, epoxy resins etc.

Difference between Thermoplastics and Thermosetting Resins

	Thermoplastic	Thermosetting Resin
1.	They are generally long chain linear polymers.	They have three dimensional crosslinked network structure joined by strong covalent bonds.
2.	On heating, they soften readily because secondary forces between the chain can be easily broken.	They do not soften on heating because cross links and strong covalent bonds retain their strength on heating. They can, however, be charred on strong heating.
3.	They can be softened, reshaped and even reused by reheating to a suitable temperature.	They can not be reshaped and reused because they do not soften on heating and retain their shape and structure.
4.	They are generally weak, soft and less brittle.	They are usually hard, strong and more brittle.
5.	They can be reclaimed from waste.	They can not be reclaimed from waste.
6.	They can dissolve in some organic solvents.	They are insoluble almost in all organic solvents because of the presence of cross linking and strong covalent bonds.

COMMERCIALLY IMPORTANT THERMOPLASTICS

1. Polyethene: Ethylene, one of the most important petrochemicals, may be polymerized by a number of methods to produce thermoplastic polymer, polyethene.



The starting material (monomer) is obtained either by dehydration of alcohol or by the cracking of petroleum. There are two varieties of polyethylenes, viz., low density polyethylene (LDPE) and high density polyethylene (HDPE).

(a) **LDPE :** Low density polyethylene is prepared by polymerizing ethylene at high pressures (1000-5000 atm) at a temperature of 250^0C in the presence of oxygen as a free radical initiator. LDPE consists of highly branched structure, which does not allow the polymeric chains to pack close together and hence results in low density (0.91 to 0.925 g/cm^3) of the polymer.

Properties:

- (i) Low-density polyethylene melts at 110 - 125^0C and is only 40-50% crystalline.
- (ii) It has low density (0.91 g/cm^3).
- (iii) It is a transparent polymer having moderate tensile strength and high toughness.
- (iv) It is chemically inert, slightly flexible and poor electrical conductor.
- (v) Its elasticity is preserved even at lower temperatures
- (vi) It is moisture resistant.

Uses:

- (i) LDPE films are mainly used for packing and wrapping frozen food, textile products etc because of their flexibility, chemical and moisture resistance.
- (ii) The non-polar nature of the polymer makes it suitable for insulation for electric wires and cables.
- (iii) Its inertness to chemicals and resistance to breakage makes it a useful material for the manufacture of 'squeeze bottles' and in many attractive containers.
- (iv) It is also used for making pipes for agricultural, irrigation and cold water plumbing and chemical plants,

(b) **HDPE:** High-density polyethylene is prepared by polymerizing ethylene at low pressure (3200-2000 atm) and at a temperature of 180-200°C in the presence of metal oxide catalysts. HDPE can also be produced by using Ziegler Natta catalyst by coordination polymerization. In this second method ethylene is heated to 60-200°C under a pressure of 6-7 atmospheres in the presence of a catalyst consisting of triethylaluminum and titanium tetrachloride (Ziegler-Natta catalyst). The polyethylene so produced consists of linear chains of polymer molecules which pack close together resulting in high density polyethylene (0.97 g/cm³).

Properties:

- (i) High density polyethylene has melting point in the range of 145-150° C and is 90% crystalline,
- (ii) It possesses density in the range of 0.941 to 0.965 g/cm³.
- (iii) It is much stiffer than LDPE and has higher tensile strength and hardness.
- (iv) It is chemical resistant and has significantly lower gas permeability.
- (v) Insulating and moisture resistant properties are similar to LDPE.
- (vi) Due to high crystallinity, it has low gas and water vapour permeability.

Uses:

- (i) HDPE is used in the manufacture of toys, crates closures, caps, buckets etc because it is more tough and stiffer than LDPE.
- (ii) Its greaseproof nature and chemical inertness makes it suitable for wrapping material for food products.
- (iii) Low gas permeability makes it suitable for domestic gas piping.
- (iv) It is also used for manufacturing bottles for milk, household chemicals and drug packaging.
- (v) It is used in the manufacture of overhead tanks for water storage.
Whenever high tensile strength and stiffness are required HDPE finds better use than LDPE.

2. Polypropylene: Polypropylene is probably the lightest known industrial polymer and can be prepared in isotactic, syndiotactic or atactic forms. The monomer propylene is obtained as a byproduct from gasoline refineries. Polypropylene (isotactic) is produced by polymerizing propylene in the presence of Ziegler-Natta catalyst [Al(C₂H₅)₃ - TiCl₄]]

Properties:

- (i) Polypropylene is stereoregular (isotactic) and highly crystalline polymer (slightly lower than HDPE).
- (ii) Isotactic polypropylene (IPP) melts at 160-170° C.
- (iii) It exhibits high stiffness, hardness and tensile strength,
- (iv) Its moisture resistance is comparable to that of polyethylene.
- (v) Polypropylene is resistant to many chemicals such as alkalies, acids and oils. But due to presence of methyl groups attached to alternate carbon atoms, it is susceptible to oxidation.
- (vi) Its electrical properties are similar to polyethylene.
- (vii) Films of polypropylene are impermeable to a number of vapours and gases.

Uses:

- (i) The higher melting point of polypropylene allows it to be used in products that must be steam sterilized such as baby bottles, flasks etc.
- (ii) Its fibres are extremely strong by weight and hence are used in making ropes and filaments; it is also used for the manufacture of drinking glasses.

- (iii) Polypropylene films are used to: packaging due to high clarity and their impermeable nature to a number of vapours and gases. It is also used for wrapping confectionary items (bread, biscuits etc) and potato chips.
- (iv) Due to high strength, rigidity and temperature and chemical resistance it is used for the production of chemical process equipments such as pipes and tanks.
- (v) It is used for the preparation of textile machinery parts (bobbins, spools, spindles etc), when blended with suitable elastomer it is widely used for making bumpers, radiator grills etc.
- (vi) It is also used as an insulating material for electrical wires.
3. **Polystyrene:** Polystyrene is also known as polyvinyl benzene. The monomer styrene is produced from ethylene and benzene. When ethylene is passed into benzene in the presence of AlCl_3 catalyst, ethyl benzene is formed, which on passing over iron oxide catalyst at high temperatures is converted into styrene.

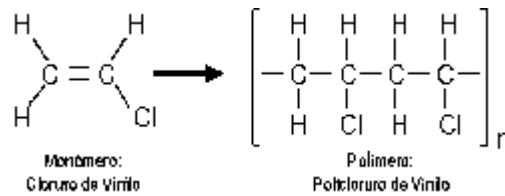
**Properties:**

- (i) Polystyrene is refractive high brilliance.
- (ii) It has got low heat distortion temperature (85°C) and hence articles made of polystyrene cannot be sterilized with steam.
- (iii) It consists generally of linear chains and is chemically inert and resistant to most alkalies, oxidising and reducing agents and acids.
- (iv) It is susceptible to photooxidative degradation, which results in brittleness and yellowing.
- (v) It has got excellent moisture resistance.
- (vi) It has good electrical insulating properties.
- (vii) It can be dyed to different colors.
- (viii) It has the unique property of transmitting light of all wavelengths through curved sections.
- (ix) Polystyrene is attacked by a number of solvents.

Uses:

- (i) Polystyrene is widely used in the manufacture of articles such as lids, jars, bottles, combs, brush handles, radio and television cabinets.
- (ii) It is used for preparing synthetic ion exchange resins in combination with p-divinyl benzene.
- (iii) An important use of polystyrene is the manufacture of foam and bead for insulation and packaging material. The polymer is foamed by using blowing agents, usually pentane.
- (iv) High-impact polystyrene (HIPS), produced by combination of polystyrene with rubbery material such as polybutadiene (SBR). Contributes to improved impact properties and is used for door liners, inner liners of refrigerators, etc.
- (v) Polystyrene copolymerized with acrylonitrile and butadiene (ABS), due to its high impact resistance and high dimensional stability at elevated temperatures is used in the manufacture of door handles, radiator grills, loudspeaker housings etc.
- (vi) Styreneacrylonitrile (SAN) copolymer is a transparent plastic with very good impact strength and is used for moulding crockery items and machine components.

4. Polyvinyl Chloride (PVC): Polyvinyl chloride is one of the cheapest and most widely used plastic globally. The monomer vinyl chloride is a gas, boiling at -14°C . Polyvinyl chloride is prepared by the polymerization of vinyl chloride by heating its water emulsion in presence of a small amount of benzoyl peroxide.



Properties:

- (i) It is colourless, odourless and non-inflammable.
- (ii) PVC is partially syndiotactic and has low crystallinity.
- (ii) PVC is thermally not very stable and degrades above 200°C , with the evolution of HCL.
- (iv) It is resistant to atmospheric oxygen, inorganic acids and alkalies.
- (v) It has greater abrasion and flex resistance compared to leather.
- (vi) It is water proof and washable
- (vii) It is less brittle and lighter when compared to glass.

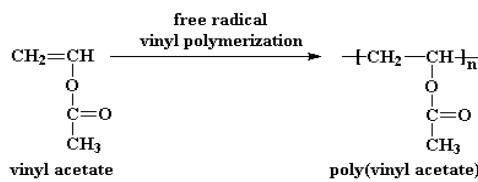
Uses: (Rigid PVC)

- (i) Rigid PVC or unplasticized PVC has superior chemical resistance and high rigidity but is brittle. It is used as sheets for tank linings, safety helmets, and mudguards.
- (ii) It is used for making window frames as a substitute for wood because the frames do not corrode.
- (iii) It is used for the manufacture of containers for detergents, cosmetics, mineral water.
- (iv) Because of its oil-resistant property, it is used for storing consumable liquids like edible oils, fruit squashes, table wine etc.

(Plasticized PVC)

- (i) Plasticized PVC is used as an insulating material.
- (ii) Because it is washable and has greater abrasion resistance therefore it is used for making ladies-handbags, bathroom curtains, kitchen upholstery.
- (iii) Used for upholstery for seats in public transport vehicles,
- (iv) In the film form it is used in membrane separation technology.

5. Polyvinylacetate (PVA): Polyvinyl acetate is produced by heating vinyl acetate in the presence of small amount of benzoyl peroxide.



Properties

- (i) Polyvinyl acetate is a solid, clear resin, and has no taste or odour and is soft sticky material.
- (ii) It is insoluble in water and is saponified with great difficulty and hence is not absorbed by the digestive system. Moreover if at all saponification takes place, the products polyvinyl alcohol and acetic acid are not harmful. Due to the above-slanted reasons, polyvinyl acetate is used as a basic material for chewing gums and surgical dressings.

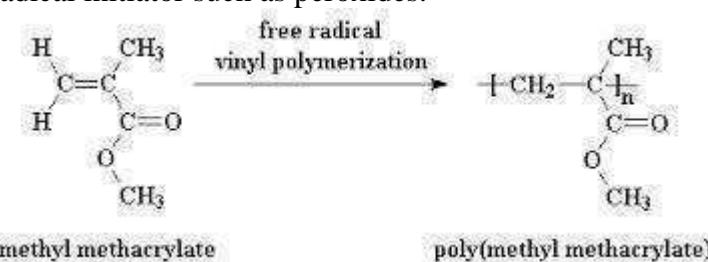
- (iii) It is resistant to atmospheric oxygen, water and chemicals.

Uses

- (i) Polyvinyl acetate is used for the manufacture of lacquers, paints and adhesives. Polyvinyl acetate adhesives are used for bonding of textile to textile, leather to leather/paper to paper, paper to textile, leather to wood etc.
 - (ii) It is used as a basic material for chewing gum and surgical dressings.
 - (iii) It is also used for finishing of textile fabrics.
 - (iv) On copolymerization with polyvinyl chloride, it is converted into a polymer having varying degree of softness. For example, vinylite, which is a copolymer of polyvinyl acetate and polyvinyl chloride, is especially useful as a surface coating material for metals.
 - (v) It is mainly used for the manufacture of polyvinyl alcohol.

Polyvinyl butyral is tough and elastic and is used for laminating safety glass. It holds the broken pieces of glass in case of accident and thus minimizes the danger from flying glass fragments.

5. Polymethylmethacrylate (PMMA): Polymethyl methacrylate (PMMA) is one of the important thermoplastic from commercial point of view, because of its outstanding optical properties. It is prepared by the polymerization of methylmethacrylate in presence of free radical initiator such as peroxides.



Properties

- (i) It is a colorless transparent plastic with an excellent outdoor life period and good strength.
 - (ii) Owing to the presence of bulky side groups, it is amorphous in nature.
 - (iii) It can be depolymerized to yield back the monomer.
 - (iv) It has high softening point of about $130\text{-}140^{\circ}\text{C}$ but it becomes rubber-like at a temperature above 65°C .
 - (v) The main feature of this plastic is its high optical transparency and capacity of transmitting light accurately, even in curved sections.
 - (vi) It has good mechanical properties but has poor scratch resistance as compared to glass.
 - (vii) It has low chemical resistance to hot acids and alkalies.
 - (viii) It has excellent weathering properties.

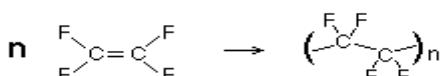
Uses

- (i) PMMA is the nearest alternative to glass as it does not shatter and can be moulded into any shape. Infact, PMMA sheets known as plexiglass or lucite are used as an alternative to glass.
 - (ii) PMMA finds major use in automotive industry (e.g. housing for rear lamps, profiles, light fittings, ceiling lighting signal-light etc).
 - (iii) It is used for making hard contact lenses due to its high transparency.

- (iv) It is used as composite materials for kitchen sinks, basins and bathtubs because it provides wide range of colour and are cheap and light.
- (v) PMMA is also used for making windscreens T. V. screens etc. The fragments formed after cracking are less sharp and hence less harmful.
- (vi) It is also used for decorative purposes such as in jewellery etc. (vii) It is also used in bone splints, artificial eyes, dentures etc.

6. Polytetraflouro Ethylene (Teflon) (PTFE): Polytetra flouroethylene is prepared by the polymerization of tetraflouoroethylcne ($\text{CF}_2=\text{CF}_2$), under high pressure in presence of beczoyl peroxide as catalyst.

The polymerization is associated with the release of large quantities of heat and if sufficient precautions are not taken, this can result in violent explosions.



Properties

- (i) PTFE has a highly regular structure and is highly crystalline.
- (ii) Due to the presence of highly electronegative fluorine atoms and regular configuration of PTFE molecule there are very strong attractive forces between different chains, which give the material extreme toughness.
- (iii) PTFE has high softening point of about 327°C . above 400°C it decomposes without changing into vapour.
- (iv) It is extremely resistant to attack by chemicals such as strong alkalies and acids. Only hot alkali metals and hot flourine can attack PTFE.
- (v) PTFE is highly dense polymer (2.1 to 2.3 g/cm) having excellent mechanical properties.
- (vi) It can be machined, punched and drilled.
- (vii) It has a very low coefficient of friction
- (viii) It has extremely good electrical insulating properties.
- (ix) It is slippery and waxy to touch.
- (x) It does not melt like other thermoplastics and hence behaves somewhat like a thermosetting polymer.
- (xi) It has non-adhesive characteristics.

Uses

- (i) Due to its chemical inertness, it is used in chemical carrying pipes, tubing's, stop-cock for burettes,
- (ii) Its extreme inertness and stability to temperature makes it a suitable material for making gaskets and seals.
- (iii) It is used as an insulating material for transformers, cables, wires, fittings etc.
- (iv) Because of its low coefficient of friction, it is used in non-lubricating bearings.
- (v) Its most important application is, as non-stick coating for Dots and pan:.. PTFE is applied as a coating over the rollers and mixing equipments used for preparing sticky materials such as adhesives and jams.
- (vi) Its fiber is used to form belts, filter cloth etc.
- (a) **Polyamides:** Polyamides are a group of polymers which contain the amide ($-\text{CONH}-$) linkage in the main polymer chain.. Polymers of this type are the synthetic

linear aliphatic polyamides, which are capable of fibre formation.

Nylon is the generic name for synthetic polyamides capable of forming fibres. Polyamides (nylons) are prepared by the polycondensation between dicarboxylic acids and diamines. Nylons are designated by two numbers, the first representing the diamine. Polyamides of commercial importance are Nylon 6.6; Nylon 6, 10, Nylon 6 etc.

- (i) Nylon 6.6: Nylon 6,6 is prepared by the polycondensation of hexamethylene diamine with adipic acid.

Properties of Nylons

- (i) Nylons are characterized by a combination of high strength, elasticity, toughness and abrasion resistance.
- (ii) They have very good moisture resistance.
- (iii) They are insoluble in common solvents.
- (iv) The molecular chains are held together to each other by hydrogen bonds having linear structure, permitting side-by-side alignment.
- (v) They are very flexible and retain their original shape after use.
- (vi) Nylons provide excellent resistance to wear and abrasion
- (vii) Due to the presence of hydrogen bonding, all nylons are water sensitive. Water acts as plasticizer, which reduces tensile strength and dimensional stability.

Uses

- (i) Nylon 6.6 is used as a plastic as well as a fibre. It's most important fibre applications include automobile tire cords, ropes, threads, cords having high tenacity and good elasticity.
- (ii) Nylon 6.6 is also used to make textile fibres for use in dresses, undergarments, socks etc.
- (iii) Nylon 6.6 being a tough plastic is used/as a substitute for metals in gears and bearings etc. Nylon 6.6 can be moulded in one piece whereas similar item in metal require assembling of several parts. Moreover, nylon bearings and gears perform quietly and need little or no lubrication.
- (iv) It is also used in making rollers, slides and door latches.
- (v) Nylon 6. 10 is not used much as synthetic fibre, but is used to manufacture articles like brushes and bristles.
- (vi) Nylon 11, is used as textile fibre.
- (vii) Due to its good barrier properties, nylon-6 film is used for packaging some oxygen sensitive food.
- (viii) Due to its high tenacity, nylon is suitably used for parachute fabric.(ix) Nylons are also used in making transmission belting, ribbons for electric-typewriters, spindle bands on textile spinning machines etc because of its high durability, (x) Nylon is also used for making insect screens because of their biological resistance.

Aromatic polyamides, often called aramids have been produced to improve the heat and flammability resistance of nylons. These polymers have very high melting points (greater than 500°C). One such nylon is Kevlar. It is prepared by polycondensation between aromatic dichloride and aromatic diamines.

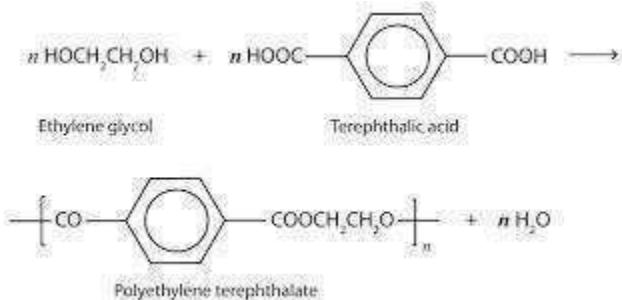
Kevlar is exceptionally strong and decomposes only above 500° C It is 5 times stronger than steel and it is used as a substitute for steel in belted radial tires as well as in bullet resistant vests. It is also used in making car parts such as tyres, brakes, clutch linings etc.

COMMERCIALLY IMPORTANT THERMOSETTING RESINS

1. Polyesters: Polyester resins are the condensation products of dicarboxylic acid with dihydroxy alcohols. The industrially most important polyester is poly (ethylene terephthalate) (PET), with the trade name Terylene or Dacron. PET can be prepared from terephthalic acid and ethylene glycol, either by direct esterification or by catalyzed ester inter change.

Nowadays, dimethyl terephthalate is used instead of terephthalic acid. This is because of the fact I thai terephthalic acid is nonvolatile and is very slightly soluble in most of the solvents. The reaction takes place in two successive stages:

The enihyiene giycoi so formed is distilled off at such high temperature.



Properties

- (i) PET is resistant to heat and moisture.
- (ii) It is strongly resistant to both weak acids and bases at room temperature but readily attacked by strong acids and bases.
- (iii) Presence of numerous polar groups results in powerful dipole attractions in it.
- (iv) It has good mechanical strength up to 175°C.
- (v) It is resistant to almost all kinds of insect attacks.
- (i) It can be dyed at 100° C or in presence of a carrier.
- (ii) The fibre has good wrinkle resistance due to good elasticity,
- (iii) PET fibres are abrasion and oxidation resistant.

Uses:

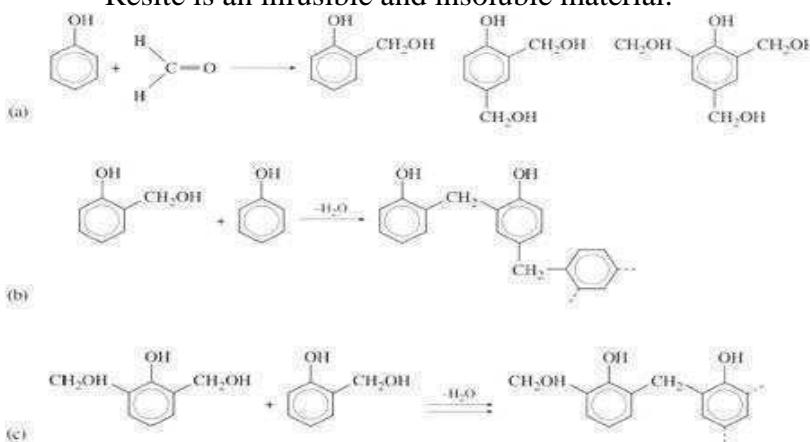
- (i) PET is mostly used for making synthetic fibres like Terylene, Dacron etc. Garments made from its fibres resist the formation of wrinkles and have many desirable properties such as warmth to touch, low moisture absorption (and hence dry rapidly), crease resistance, shape retention, and mildew proof. Moreover, these may be blended with other fibres, such as wool end cotton.
- (ii) PET is also useful for making films because of its high strength, resistance to tearing durability, transparency etc. PET is used in making the transparencies of overhead projector.
- (iii) Due to its impermeability to a number of organic and inorganic gases it is used in the manufacture of plastic bottles for carbonated beverages (barrier to CO₂ and water)
- (iv) It is also used to make magnetic recording tapes.
- (v) Due to their high dielectric strength they are used for electrical insulation.
- (vi) Glass reinforced PET is used for the manufacture of housings-for coffee machines, car heaters, toasters.

- 2. Phenolic Resins:** Phenol formaldehyde resins are the most important and extensively used synthetic plastics and perhaps the oldest synthetic plastics. In these resins one of the monomers is always a phenol molecule. One of the most important member of this class is

phenol formaldehyde resin.

The phenol formaldehyde reaction is catalyzed by alkalies and acids. The nature of the product depends on several factors, the major one being the nature of the catalyst and the proportion of the reactants (phenol formaldehyde ratio). The formation of phenol formaldehyde resin comprises of the following steps:

- (i) **Methyolation:** When phenol and formaldehyde react together, the first step is the entry of methyloL (CH_2OH) groups in ortho and para positions to the hydroxy group. The reaction takes place in the presence of acid or alkali. Depending upon phenol formaldehyde ratio, various phenol-alcohols may be formed.
- (ii) **Novolac formation:** Depending upon the ratio of phenol formaldehyde (P/F) different resins, namely novolac and resol resins are obtained. In the presence of acid catalyst, when the P/F ratio is greater than unity, the methyloL derivatives condense with phenol to form a linear polymer with little methyloL groups. The product is thermoplastic in nature and is known as novolac.
- (iii) **Resole formation:** In the presence of alkaline catalysts with P/F ratio less than 3, the methyloL phenols condense to form linear structure called resoles. The resoles are soluble in the reaction mixture and have excess of methyloL groups capable of further reaction during continued heating. Resols, therefore have limited shelf-life. However, the reaction may be stopped at the desired step by cooling.
- (iv) **Resite formation:** On further heating, resoleS are transformed into a three dimensional polymer, calied resite, due to the presence of active groups and mobile hydrogens. Resite is an infusible and insoluble material.



Properties

- (i) Phenol formaldehyde moulding resins have excellent heat resistance.
- (ii) These have high dimensional stability.
- (iii) To improve the properties further, fillers such as wood, flour, mica, asbestos etc are added to them.
- (iv) Phenolic resins have good dielectric properties.
- (v) These have remarkable adhesive properties and bonding strength.
- (vi) They are however attacked by alkalies, because of the presence of free hydroxy groups.
- (vii) They are hard, rigid and scratch-resistant.

Uses

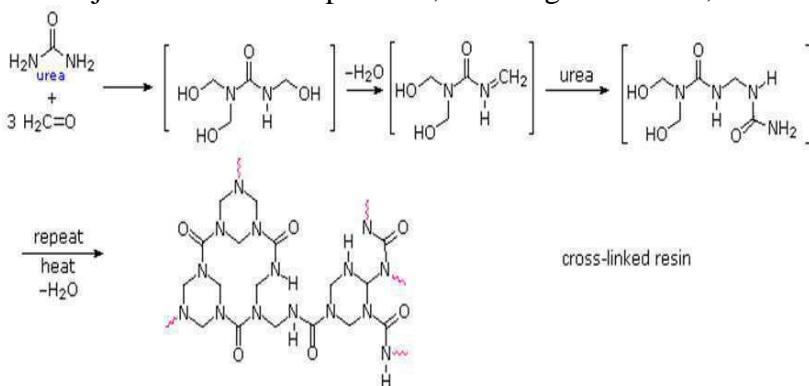
Phenolic resins have wide applications.

- (i) The main use of phenolic resin is in moulding applications. They are widely used in making telephone parts, cabinets for radio, television and automobile parts

- (ii) Due to their adhesive properties and bonding strength, they are used for producing brake linings, abrasive wheels and sand paper.
- (iii) They are used for making electric insulator parts like switches, plugs, switch boards, heater-handles etc,
- (iv) Phenolic resins are used for impregnating paper, wood and other fillers.
- (v) These are also used in varnishes, paints and protective coatings.
- (vi) These are used in the production of ion exchange resin for water softening.

4. Amine Resins: Amino resins or amino plasts are prepared by the reactions of formaldehyde with nitrogen compounds such as urea, melamine and other amino compounds. Because of their attractive light colours, these resins have great commercial importance.

- (a) **Urea formaldehyde resins:** This is prepared by the condensation reaction between urea and formaldehyde in neutral or alkaline conditions. The first products formed during the formation of resin are monomethylol and dimethylol ureas.
- (ii) Polymerization can take place from mono or dimethylol urea, or possibly through both, with the formation of long chains.
- (iii) Thermosetting of the resin takes place during the moulding process. For moulding, the methylol derivatives are compound with fillers, plasticizers, pigments etc. and then are subjected to heat and pressure, resulting into a hard, infusible product.



Properties

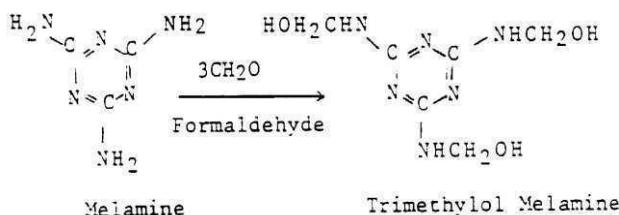
- (i) The urea formaldehyde resins are light in colour and enable the production of light or pastel coloured objects.
- (ii) Their heat and moisture resistance are lower than those of phenolic resins.
- (iii) They have good electrical insulating properties.
- (iv) They are resistant to oxidation to oil, grease and weak acids.
- (v) They are hard, resist abrasion and scratching.
- (vi) They have good adhesive properties.

Uses

- (i) Urea-formaldehyde resins are widely used to make moulded articles like vacuum flask cups jugs buttons, bottle caps etc where the darker colour of phenolics may be objectionable.
- (ii) They are also used for adhesive applications for the production of plywood and laminating.
- (ii) They are used for the manufacture of coated abrasive paper and binder for foundry

cores.

- (iv) They are used for the manufacture of cation exchange resins.
 - (v) These also find use in the manufacture of electrical switches, plugs and insulating foams,
 - (vi) Their applications also include the treatment of textile fibers for improving their shrink and crease resistance.
- (b) **Melamine-formaldehyde resin:** Melamine formaldehyde resins are another class of amino resins. They are produced by the condensation of melamine (2, 4, 6 triamino – 1, 3, 5 triazine), trimer of dicyandiamide with formaldehyde, exactly in the same manner as in the formation of urea formaldehyde resin. The reaction takes place under mild alkaline or neutral conditions.



Properties:

- (i) The melamine resins have better hardness, heat resistance, and moisture resistance than the urea - formaldehyde resins. However these are expensive. Rest of the properties is similar to urea-formaldehyde resins.

Uses

- (i) Typical applications of urea formaldehyde resins include the production of decorative plastic ware, laminated worktops.
- (ii) They are used in electrical fittings.
- (iii) They are used for resin bonded grinding wheels.
- (iv) They are used for textile finishing providing crease resistance, stiffness, water repellency etc.
- (v) They are also used in automotive finishes.
- (vi) Melamine resins are also used as a tanning material for leather. The presence of small amount of this resin in the tanning material improves the strength of the leather.
- (vii) They are also used for the preparation of lacquer

RUBBERS (ELASTOMERS)

It was discussed in the beginning of the chapter that depending on the ultimate form and use a polymer can be classified into two broad categories: Plastics and Elastomers (rubbers).

Elastomers are the polymers which are capable of being stretched rapidly at least 150 percent of their original length without breaking and return to their original shape on release of stress. For example, rubber band is a typical elastomeric material. When stretched, it is elongated to many times its length. But when the force is released it gets back to its original length.

The properties of typical elastomers are defined by the following:

- They must stretch rapidly and considerably when stretched, reaching high elongations (500-1000%).
- They must exhibit high tensile strength and high stiffness when fully stretched.
- They must snap back to their original length once the stretching force is withdrawn or the

stress is released.

These criteria are met with the materials which:

- Have high molecular mass (high polymers).
- Are above their glass transition temperature.
- Are amorphous.
- Contain a network of crosslinks to restrain gross mobility of its chains.

Columbus on this voyage to America found the native of West Indies were playing with the solid mass collected from the exudate of a tree. Years later, Joseph Priestly showed that this solid mass could rub off pencil marks from paper and hence the material was called rubber. The main source of natural rubber is the species of tree known as 'Hevea Basiliensis' even today, which account for multibillion-dollar rubber industry. Rubber or elastomers can be obtained naturally or synthetically. Consequently we have natural rubber and synthetic rubber.

NATURAL RUBBER

Natural rubber of nearly same characteristics occur in the inner bark and to a lesser extent in the leaves and roots of a number of tropical plants. But the latex from the tree 'Hevea Basiliensis' is proved to be the main source of rubber. Latex is an emulsion of poly hydrocarbon droplets in an aqueous solution and looks like milk. The negatively charged particles of polyhydrocarbons, a colloidal dispersion, stabilize the emulsion.

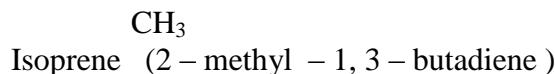
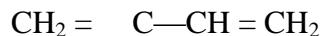
Processing

The latex is collected by making small incisions in the trees in such a manner so as to allow the latex to accumulate in small cups. The percentage of rubber in it is 25-35%. A fully grown tree gives about 5 gm of rubber in a day.

The average composition of the latex is:

Water = 60%; Rubber = 35%; Proteins, enzymes and nucleic acid = 3%; fatty acids and esters = 1% and inorganic salts = 0.5%.

In these trees, the rubber is formed from polymerization of isoprene i.e. (2-methyl-1, 3-butadiene) to produce poly-(cis)isoprene by a biochemical reaction in which a particular type of enzyme acts as a catalyst.



The latex is first sieved to remove the impurities such as leaves, bark etc. From 25-35% dry rubber content the latex is diluted to contain about 15-20% of rubber content. The rubber is then separated from the latex by the process of coagulation. The acetic acid or formic acid is added as coagulants and the rubber is coagulated to soft white mass. About 1 g of acetic acid/formic acid is added per 2G0 kg of rubber material. This coagulated mass is the crude rubber and consists of 90-95% rubber, 2-4% proteins and 1-2% resins. The crude rubber is washed and treated for producing the following types of rubber depending upon the process.

1. Crepe Rubber: Crepe rubber is prepared by adding a small amount of sodium bisulphite (NaHSO_3) to the coagulated mass to bleach the rubber. Then the rubber is passed through a creping machine. The creping machine is rolling machine which consists of two rollers having longitudinal grooves. Thus, after passing through it, the spongy rubber is converted into a sheet having a rough surface which resembles the crepe paper and hence the name.

2. Smoked Rubber: The smoked rubber is produced by carrying out the coagulation of latex in the tanks provided with vertical grooves on sides and fitted with metal plates which run

across the width of the tank. First the diluted latex is poured into the tanks, with the metal plates removed. Then the coagulant is added (acetic or formic acid) and the plates are inserted. The tanks are kept undisturbed for 15-16 hours. The tough slabs are formed. These slabs are passed through a series of rolling machines and finally rubber sheets having ribbed pattern are obtained. The ribbed surface prevents the sticking of the rubber with each other. These sheets are then dried in smoke houses at about 50°C in the smoke obtained from burning wood or coconut shells. The smoked rubber so obtained is translucent and ambre coloured.

GUTTA PERCHA

Gutta percha is another variety of natural rubber; it is obtained from the mature leaves of dichopsis gutta and palagum gutta trees. The hevea rubber and gutta percha, both are composed of isoprene units with structures shown in the figure? But **natural rubber** is cis-isomer where a bending back of successive isoprene units take place giving the molecule a coiled structure. The gutta percha, on the other hand is a trans-isomer which consist of straight chains, giving a rod like structure to the molecule. The rod-like structure results into close packing of adjacent chains. Thus gutta percha is more crystalline.

Processing of Gutta Percha

The mature leaves obtained from the trees Dichopsis gutta or palagum gutta are carefully ground in mills and are then treated with water maintained at 70°C for about half an hour when the gutta percha is extracted in water. The solution is dropped into cold water and gutta percha floats on the surface. Alternatively, gutta percha can be recovered by solvent extraction.

Properties of natural (Hevea)rubber and gutta percha:

Natural Rubber (Hevea rubber)	Gutta Percha
1. Natural rubber is highly soft and elastic material	Gutta percha is a hard thermoplastic material at room temperature but it softens and becomes tacky at about 100°C
2. It is soluble in carbon disuphid- and petrol.	It is soluble in petrol,
3. It Is cis polyisoprene	It is transpolyisoprene.

Drawbacks of Raw Rubber: Raw rubber has a number of drawbacks when put to use. Some of the drawbacks are:

- (i) Crude rubber is soft at high temperatures but becomes brittle at low temperatures. So it can be used in a limited temperature range.
- (ii) It is too weak with a tensile strength of 200 kg/cm^2
- (iii) It is not resistant to mineral oils, organic solvents.
- (iv) It has large water absorption capacity.
- (v) Due to oxidation in air, it deteriorates and hence its durability is decreased.
- (vi) It is also attacked by the oxidizing agents like nitric acid, Conc. sulphuric acid, chromic acid etc.
- (v) It undergoes permanent deformation when stretched. This may be due to stretching the individual chains slip past each other and get separated resulting in the breakage of the material.
- (vi) However the above drawbacks can be overcome by improving the properties by the addition of suitable materials and further heat treatments.

Some of the processes for improving the properties of rubber are discussed below:

PROCESSES FOR IMPROVING THE PROPERTIES OF RUBBER

The desired properties in rubber can be obtained by compounding it with certain chemicals. The compounding of rubber is generally facilitated by a process called mastication.

Mastication: When rubber is subjected to mechanical stresses arising from say, high-speed stirring or milling, it undergoes considerable molecular degradation or fragmentation. The process is called mastication. The mastication converts hard and tough rubber into a soft, supple and semisolid mass. The $\text{CH}_2 - \text{CH}_2$ links between the isoprene units in the rubber are weak and hence break when subjected to mechanical stresses. The net effect is that the big molecule is broken into small fragments. After mastication is complete the compounding agents are added to the rubber.

Compounding: The process involving the incorporation of chemicals such as plasticizers, vulcanizing agents, stabilizers, fillers etc., into the rubber industry was discovered by Charles Goodyear in 1839. the process by which a network of crosslinks is introduced into an elastomers is called vulcanization. Charles Goodyear used sulphur to crosslink polyisoprene molecules in natural rubber. The added sulphur combines chemically at the double bonds bringing about changes in its properties profoundly. Vulcanization transforms elastomers which are a weak thermoplastic mass of no use into a strong, elastic and tough rubber. The tensile strength, stiffness, durability, resistance to changes in temperature, elasticity and chemical resistance of a vulcanized rubber are increased manifolds as compared to unvulcanized rubber.

The introduction of sulphur links in between the polymer chains restrict the intermolecular movement which brings about stiffening of rubber.

The extent of stiffness of vulcanized rubber depends upon the amount of sulphur added. For example, a rubber may contain 3 to 5% sulphur whereas Vulcanite or ebonite (hard rubber) contains about 32% sulphur. The extensive Crosslinking in hard rubber makes it so stiffer tat there is practically no elasticity in it. Fig. shows the effect of vulcanization on rubber. Ordinary rubber molecules are bent and convoluted (a) When stretched the molecules can slip past one another. But vulcanization introduces crosslinks between these and hence on stretching the molecular chains cannot slip past each other.

(a)	(b)	(c)
Unvulcanized Rubber	Vulcanized Rubber (Unstretched)	Vulcanized Rubber (Stretched)

Chemistry of Vulcanization

The addition of sulphur develops crosslinks, thereby improving the processibility and utility of rubber. But the mechanism of vulcanization is quite complex and is still not known in detail. Both free radical and ionic intermediates have been postulated. Sulphur exists as S_8 i.e. cyclo octasulphur. The probable mechanism can be:

- Sulphur (S_8) dissociates into $\text{S}^+ \text{S}^-$

$$\text{S}_8 \longrightarrow \text{S}^+ \text{S}^-$$

Where m and n are the numbers of sulphur atoms.

- The sulphur cation Sm^+ reacts with a diene molecule of rubber to give I.
- This cation adds to other polymer molecule to give II and carbonium ion III
- The arlylic carbonium ion reacts with S_8 to form IV which adds to polydiene to form sulphur bridges (V). This can further react with S_8 and the cycle goes on

It must be emphasized that neither heat nor sulphur is essential for the vulcanization process. Vulcanization can even be carried out by number of other substances. Hence vulcanization can be of two types:

- (a) **Sulphur Vulcanization:** As stated earlier, in this process the rubber is heated with sulphur or CS_2 .

(b) **Non-sulphur Vulcanization:** Non-sulphur vulcanization is achieved through a free radical.

- Free radical can be generated from a curing agent such as benzoyl peroxide or as a step in the oxidative degradation of rubber.
- This free radical abstracts a hydrogen atom from one of the –methylene groups.
- The rubber free radical so formed attacks a double bond in an adjacent polymer chain resulting into crosslink (vulcanization) and regeneration of a free radical.

This cross linked free radical will further attacks more C=C bonds to form a large number of cross linkages leading to vulcanization.

Further, the compounds such as Se, Te, ZnO, benzoyl, chloride, trinitrobenzene, etc., are also used as vulcanizing agents.

Advantages of Vulcanization

Vulcanized rubber has got ideal elastomeric properties which increase its utility manifolds. The main advantages of vulcanized rubber are:

1. **Tensile strength:** Tensile strength is the measure of the ability of a polymer to withstand pulling stresses. Vulcanized rubber has got good tensile strength and extensibility. The tensile of the raw rubber is 20 kg/cm^2 , which increases to 2000 kg/cm^2 on vulcanization.
2. **Resilience:** It is the property of a elastomers to return to its original shape on the release of stress. Resilience of vulcanized rubber improves manifolds as compared to raw rubber.
3. **Water Absorption Capacity:** The water absorption capacity of vulcanized rubber is greatly reduced.
4. Vulcanized rubber is resistant to organic solvents such as benzene CCl_4 , fats and oils. Though it swells in these solvents but to a lesser extent than raw rubber.
5. **Insulation Properties:** Vulcanized rubber has better insulation properties than raw rubber. Raw rubber vulcanized with 32% sulphur (ebonite) has extremely good insulating properties.
6. **Elasticity:** The elasticity of vulcanized rubber is greatly reduced. The elasticity of the vulcanized rubber depends upon the extent of vulcanization. If 30-50% of sulphur is present in vulcanized rubber then a rigid, non-elastomeric plastic (ebonite or hard rubber) is formed.
7. The working temperature range for vulcanized rubber is -40°C to 100°C as compared to raw rubber, which has working temperature range of 10°C to 60°C .
8. It is resistant to abrasion, ageing and reactivity with oxygen and ozone.

A comparison of raw rubber and vulcanized rubber is given in the following table:

Table: Comparison of properties of Raw, Vulcanized Rubber, Reinforced Vulcanized Rubber.

Property	Raw Rubber	Vercanized Rubber	Reinforced Valcanized Rubber
1. Tensile strength (psi)	300	3000	4500
2. Elongationat break (%)	1200	800	600
3. Elasticity	Very high	Low, depends upon extent of vulcanization	Very low, depends upon extent of vulcanization and reinforcement
4. Water absorption	Large	Small	Small
5. Solvent resistance (hydrocarbons)	Sluble	Swells only	Swells only
6. Chemical resistance	Poor	Much better	Much better

Other Compounding Techniques

Though vulcanization improves the properties of natural rubber to a great extent but even vulcanized rubber do not exhibit satisfactory properties for various utility items. These propenies can be enhanced by other compounding techniques discussed below:

(i) **Reinforcing Fillers:** To give strength and rigidity to the rubber, fillers are added. Fillers are of two types: (a) inert fillers, such as clay which make the rubber mixture easy to handle before vulcanization, but have no effect on the physical properties, and (b) reinforcing fillers which greatly improve the properties of rubber or vulcanized rubber.

Common fillers include carbon black, zinc oxide, magnesium carbonate, barium, sulphate, calcium carbonate etc. Carbon black is remarkable reinforcing filler for both natural and synthetic rubbers. Table gives the comparison of raw rubber and non-reinforced vulcanized rubber and reinforced vulcanized rubber. In vulcanization, though the long range movement of the polymer molecules is restricted, still movement of local chains is high. Reinforcement of vulcanized rubber restricts this local freedom of movement, making the polymer stiffer and tougher.

(ii) **Plasticizers or Softeners:** Plasticizers are added to rubber to improve flow and therefore, processability-plasticizers also reduce the brittleness of the product. The common plasticizers used are hydrocarbon oils, resins, waxes, etc. The oil serves as a plasticizer and a softener by reducing the melt viscosity of the rubber to the levels required Compounding.

(iii) **Antioxidants:** Rubbers that retain double bonds in their vulcanized structure are sensitive to heat, light and particularly oxygen. Thus rubber deteriorates in presence of light and air due to oxidation and as a result becomes soft and degrades. Antioxidants" are substances that protect the polymer by preventing oxidation. In many cases these also prevent undesirable reactions promoted by oxygen. Common antioxidant is complex organic amines like phenyl p-naphthylamine, phenolic substances and phosphates. The structures of two common antioxidants are given below:

Antioxidants

(iv) **Accelerators:** Vulcanization takes place by heat and in presence of sulphur, but the process is very slow. The process can be speeded up by addition of small amounts of organic or inorganic compounds known as accelerators. The inorganic accelerators include lime, magnesia and white lead whereas organic accelerators are complex organic compounds such as aldehydes, amines, thicarbonates and 2-mercaptopol benzothiozole. Sometimes accelerators require the presence of still other chemicals known as activators or promoters. These activators are usually metallic oxides such as zinc oxide. The best combination for efficient vulcanization include sulphur, an organic accelerator, a metallic oxide and a soap.

(v) **Colouring Agents:** Several dyes and pigments are used to impart colours to elastomers. Common colourants for polymers include:

Appearance	Chemical used
White	TiO ₂ , BaSO ₄
Yellow	Lead Chromate
Red	Ferric oxide
Crimson	Antimony sulphides
Blue	Ultramarine
Green	Chromium trioxide
Silver metallic effect	Flake Aluminim
Pearlescent	Lead carbonate and mica.

The oil-soluble dyes or which have small particle size and refractive index near that of the

plastic (organic pigments such as phthalocyanins) are added to get transparent coloured products from transparent plastics. The inorganic pigments impart opacity to the plastic.

SYNTHETIC RUBBERS

The polymers possessing similar physical properties as that of natural rubber are called synthetic rubbers. As natural rubber suffers from many disadvantages, attempts to find out a substitute for natural rubber began very early.

Greville William (1860) was first to isolate isoprene as a decomposition product by heating rubber to a high temperature. This gave a clue that rubber can be synthesized by the polymerization of monomers like isoprene. By the end of 19th century the fact was confirmed when polymerization of isoprene yielded rubber like product.

Since then attempts were on to prepare synthetic rubber by other monomers as well. Synthetic rubbers have been produced by the polymerization of a large number of conjugated dienes resembling isoprene. The foremost of all is, Buna Rubber (polymer of butadiene) which was prepared by the action of sodium on butadiene. That is why synthetic rubber is known as Buna rubber—'Bu' for Butadiene and 'Na' for sodium. However, the product prepared had poor tensile strength, and ageing characteristics. Improvements in properties were made by developing copolymers of butadiene. Several types of synthetic rubber are available today and some of them are described below:

1. SBR or Buna-s Rubber

Styrene-butadiene rubber (SBR) is the first commercially made synthetic rubber. It is a copolymer obtained by the free-radical polymerization of two monomers, styrene and butadiene. The two components are allowed to react in a mixing vessel containing an aqueous solution of an emulsifying agent in presence of free radical initiator such as. cumene hydroperoxide or p-menthane hydroperoxide.

According to the temperature used, two processes have been developed, the hot process (at 50°C) and the cold process. In the cold process the temperature is kept in the range -18°C to 5°C. The rubber is known as cold rubber. Cold process is more efficient than the hot process because chain length can be controlled more efficiently. The SBR copolymer contains about 75% butadiene and 25% styrene and is a random copolymer. Both type of rubbers 'cold' as well as 'warm rubbers' can be vulcanized by conventional methods using sulphur. A reinforcing filer is added to achieve good physical properties.

Properties

- (i) SBR possesses high abrasion loses resistance and high load bearing capacity,
- (ii) At elevated temperatures SBR almost 2/3 of their tensile strength whereas natural rubber only loses 1/4th.
- (iii) Its resilience is poorer than that of natural rubber, which restricts its use in tire industry.
- (iv) The weather ability of SBR is better than that of natural rubber.
- (v) 'Cold' SBR is superior to the standard product because it contains less chain branching and crosslinking.
- (vi) SBR gets easily oxidized in presence of traces of ozone present in atmosphere.
- (vii) It swells in oils and solvents.

Uses

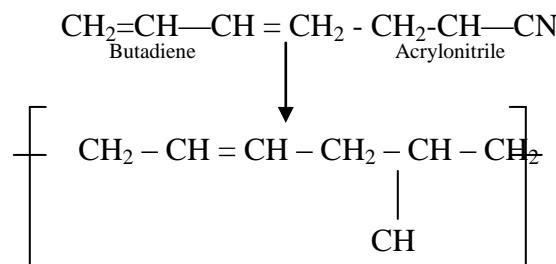
Of all the synthetic elastomers; SBR enjoys the major share of the market. SBR has the

following applications:

- (i) The principal use of SBR is in the production of tires and tire products which are typically blends of both natural and synthetic rubbers. SBR is used for lighter duty tyres but is inferior to the natural rubber in its use for truck tyres.
 - (ii) For many mechanical goods, it is superior to natural rubber and is used because of its easier processing.
 - (iii) It is used for making hose, shoe soles, gaskets, floor tiles, carpet backing, tank linings etc.
 - (iv) It is widely used for electrical insulation, although its properties are not good as those of butyl rubber.

2. Buna-N or Nitrile Rubber (NBR) or GR-A

The nitrile rubbers are polymers of butadiene and acrylonitrile in the ratio of 75% and 25% respectively. !^{f} is prepared by emulsion polymerization using cumene hydroperoxide as initiator. Hydrogen peroxide and ferros sulphate can also be used as initiators. Vulcaniza ion is done by conventional means by adding sulphur.



- (i) Nitrile rubbers are noted for their oil resistance. However the oil resistance varies with composition. If the acrylonitrile content is between 18% then the rubber has fair oil resistance and if it is more than 40% then it is extremely oil resistant.
 - (ii) It possesses excellent resistance to heat, sunlight, acids and salts.
 - (iii) Because of the presence of cyano groups (-CN). it is less resistant to alkalies than natural rubber.
 - (iv) it has good tensile strength and abrasion resistance.
 - (v) The rubbers are less resilient than natural rubbers.
 - (vi) It has good heat resistance and if properly protected by antioxidants, it shows resistance to oxidative degradation as well.

Uses

- (i) Nitrile rubber is used primarily for its oil resistance. These are extensively used for fuel tanks, gasoline hoses, gaskets, printing rollers and high altitude aircraft components.
 - (ii) They can be used in contact with water and antifreeze solutions.
 - (iii) They are also used in the form of latex for impregnating paper, textile and leather.
 - (iv) It can also be used as blend component in tyre manufacture.

3. Neoprene or GR-M Rubber

This is also known as polychloroprene or polychlorobutadiene or duprene. Neoprene was the first synthetic rubber developed in the United States.

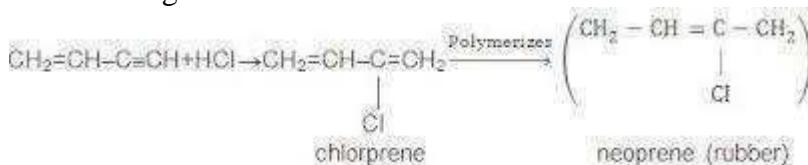
Preparation: Neoprene or polychloroprene is prepared by polymerization of

chloroprene, which is 2-chloro-1, 3 butadiene.

The preferred configuration produced during the polymerization is trans 1,4-polychloroprene. The polymerization of chloroprene takes place through emulsion polymerization.

The starting material, chloroprene is manufactured by the addition of hydrogen chloride to vinyl acetylene, which in turn is prepared by dimerization of acetylene.

The vulcanization of neoprene is different from that of other elastomers in that it can be vulcanized by heat alone. Although some types are polymerized in presence of sulphur which introduces crosslinks in the polymer. Neoprene can also be vulcanized using zinc oxide or magnesium oxide.



Properties: Neoprene is closely related to the natural rubber and can replace rubber because similarly in the properties of the two. Some of the properties of Neoprene are:

- (i) As H atom in the natural rubber is replaced by Cl atom in Neoprene, Neoprene has superior resistance to vegetable and mineral oils.
 - (ii) It is superior to natural rubber as far as ageing and effects of high temperatures are concerned.
 - (iii) Neoprene is soluble in polar solvents due to the presence of chloro-group.
 - (iv) Vulcanized neoprenes have high tensile strength (3500-4000 psi) in the absence protected.
 - (v) Vulcanized neoprene is extremely resistant to oxidative degradation if suitably protected.
 - (vi) It has good weathering and ozone resistance.
 - (vii) It has got excellent oil resistance, though inferior to Nitrile rubber but better than natural rubber or SBR.

Uses

Neoprene is used for:

- (i) It is used in oil resistant wire and cable coatings, industrial hoses, shoe holes.
 - (ii) Its latex is used for the manufacture of gloves and coated fabrics.
 - (iii) It is used for making tubings for carrying corrosive gases and oils.
 - (iv) It is used for the manufacture of sponges, conveyor belts, linings of reaction vessels etc.

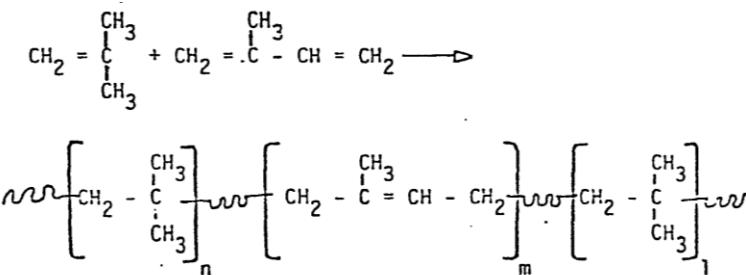
4. Butyl Rubber (GR-I)

Butyl rubber belongs to the class of elastomers which do not have unsaturated chain structure. However, a small amount of diene monomer that provides a site for vulcanization is included by means of copolymerization.

Preparation

Butyl rubber or polyisobutylene are prepared by the copolymerization of isobutylene with a small amount (1.5 to 4.5%) of isoprene. Isoprene is added to render them vulcanizable. It is prepared by low temperature cationic polymerization using Friedal-crafts catalyst at

-80°C.



Properties

The absence or low concentrations of double bonds in the main chain of butyl rubber results in high resistance to attack by a number of chemicals. It has the following properties.

- (i) Polyisobutylene or butyl rubber is amorphous under normal conditions but crystallize on stretching.
- (ii) It exhibits outstanding resistance to attack by oxygen and ozone due to low concentration of double bonds. It is less sensitive to oxidative ageing as compared to other elastomers except silicones.
- (iii) Butyl rubber becomes soft on oxidative degradation.
- (iv) It exhibits low gas permeability due to close packed linear paraffinic chains.
- (v) It has high ozone and acid resistance.
- (vi) Due to very low unsaturation, it cannot be hardened much on vulcanization.
- (vii) Unstabilized polyisobutylenes are degraded by light or heat to sticky low-molecular weight products. Therefore usual rubber antioxidants should be added to stabilize the polymer.
- (viii) It has excellent insulating properties as predicted from its non-polar, saturated nature.
- (ix) It is soluble in hydrocarbons such as benzene.
- (x) The tear resistance of butyl rubber is quite good and is retained well at high temperature and for long times, in contrast to natural rubber.

Uses

- (i) Butyl rubber is used for inner tubes for tires because of its impermeability to gases.
- (ii) It is used for wire and cable insulation.
- (iii) It is used for linings of tanks, hoses etc.

5. Polysulphide Rubber (Thicol or GR-P)

Polysulphide Rubbers are the condensation products of ethylene dihalides and alkali sulphides. These are linear condensation polymers.

Preparation

Polysulphide rubber is obtained by the reaction between sodium polysulphide and ethylene dichloride. Thus sulphur forms the part of the polymer chain. During the reaction NaCl gets eliminated. Vulcanization cannot be achieved by sulphur because polymer chain is not unsaturated. Hence it is not a hard rubber. However vulcanization can be achieved by reacting with ZnO which reacts with sulphur links forcing crosslinks.

Properties

The properties of the materials depend on the length of the aliphatic groups and the

number of sulphur atoms present in it. With four sulphur per monomer, the product is rubbery, whereas if only two sulphur atoms are present per monomer, the polymer do not behave as elastomer. In that case at least four methylene groups are needed in the dihalide to induce elastomeric properties to the products. Some of the important properties are:

- (i) Thiokol is resistant to swelling.
- (ii) It is resistant to oils and solvents. Fuel oils, lubricating oils, gasoline and kerosene have no effect on Thiokol.
- (iii) Thiokol films are impermeable to gases to a large extent.
- (iv) It cannot be vulcanized and hence does not form hard rubber.

Thiokol rubbers from many disadvantages:

- (i) It has disagreeable odour (mercaptan odour).
- (ii) It has poor heat resistance, poor abrasion resistance and low tensile strength.
- (iii) It tends to lose its shape under continuous pressure.

However, recent developments of Thiokol have overcome these disadvantages.

Uses

- (i) Polysulphides or Thiokol is used to make sealants, gaskets, balloons, fabric coating.
- (ii) Due to low permeability it is used to make gasoline hoses.
- (iii) Liquid Thiokol can be cured at room temperature to tough solvent-resistant rubbers. These liquid compounds are used as gasoline tank sealants and liners for aircraft and in variety of other sealing applications.
- (iv) The liquid polymers, when combined with oxidizers burn with great intensity and generate large volume of gas. Thus it can be used in solid-fuel rocket propellants.

6. Polyurethane (PU)

Polyurethanes are polymers containing the group (carbonate group) these are known as urethane linkages

Preparation: The polyurethanes are formed through the reaction diisocyanate with the compound containing at least two active hydrogen atoms such as diols or diamines. The reaction of diisocyanate with glycol is represented as:

The properties of the final product depend upon the type of monomers and their chain length. A wide variety of polymers can be prepared ranging from rigid to elastic polymers. Crosslinking takes place through isocyanate groups which are capable of reacting with hydrogen atoms attached to the polymers leading to vulcanization. The polyurethanes can be produced in the form of fibers, elastomers or foams.

The foamed plastic can be obtained by blowing air/gas through molten polyurethane. This foamed plastic is also known as thermocol.

Note: Thermocol is also prepared by blowing air/gas through polystyrene.

Alternatively, for the production of foamed polyurethane little water is mixed with the reaction mixture, which reacts with isocyanate group to evolve CO_2 . The CO_2 produced foams the polymer as it sets.



Properties

- (i) Polyurethanes have high strength, good resistance to gas and aromatic hydrocarbons.
- (ii) They have excellent resistance to ozone and oxygen but are susceptible to microbial attack.
- (iii) The polyurethanes can be prepared in different forms varying from hard coatings to

flexible rubbers and rigid foams each having different properties.

- Coatings:** Coatings are hard, glassy and solvent resistant. They have good flexibility and impact resistance.
- Elastomers:** Elastomers have excellent abrasion resistance and are resistant to oils, greases and solvents.
- Fibers:** Fibers are obtained by reacting di-isocyanate with long chain glycols which are usually polyesters or polyethers.
- Foams:** Foams can be flexible or tough depending upon the polymer and the type of Crosslinking produced. They are light, tough and resistant to heat.

Uses

- (i) The polyurethane foams have found the widest use. The flexible foams are used as upholstery material for furniture, insulation and crash pads.
- (ii) The rigid polyurethane foams are used as core material as insulation cover in refrigerated appliances and vehicles.
- (iii) Flexible polyurethanes can also be used for coating rubber articles to give them additional resistance to abrasion, such as in gymnasium and dance floors.
- (iv) Polyurethane resin is used to insulate the wires polyurethane coated wires can be soldered directly.
- (v) Polyurethane fibers with unusually high elasticity are used in foundation garments and swim wear. They have replaced rubber latex thread in this use.
- (vi) Flexible urethane foams are used as cushions for furniture and automobiles and have replaced rubber foams in these applications because of improved strength, lower density and easier fabrication.
- (vii) Polyurethanes are also used in making mattresses and pillows.
- (viii) As they are radiation resistant, they are also used in encapsulation for integrated-circuit devices.
- (ix) Its applications in medical field include its use in dental materials, artificial heart and skin.

7. Chlorosulphonated Polyethylene Rubber (Hypalon)

Preparation: When polyethylene is treated with a mixture of chlorine and sulphur dioxide, chlorosulphonated polyethylene is formed. The chlorosulphonation can be carried out either on the solid material or in solutions.

Properties

The chlorosulphonylation of ethane provides two modifications:

- (a) The chlorine atoms break the regularity of polyethylene chain structure changing the crystalline polyethylene to an elastomer.



RGPVNOTES.IN

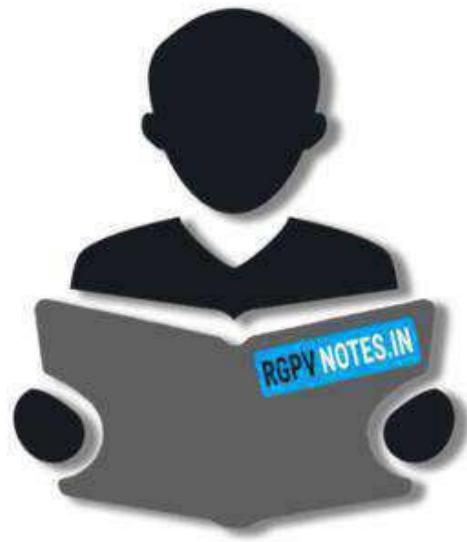
We hope you find these notes useful.

You can get previous year question papers at
<https://qp.rgpvnotes.in> .

If you have any queries or you want to submit your
study notes please write us at
rgpvnotes.in@gmail.com



LIKE & FOLLOW US ON FACEBOOK
facebook.com/rgpvnotes.in



RGPVNOTES.IN

Program : **B.Tech**

Subject Name: **Engineering Graphics**

Subject Code: **BT-105**

Semester: **1st**



LIKE & FOLLOW US ON FACEBOOK
facebook.com/rgpvnotes.in

Module - IV

Sections and Sectional Views of Right Angular Solids covering, Prism, Cylinder, Pyramid, Cone – Auxiliary Views; Development of surfaces of Right Regular Solids - Prism, Pyramid, Cylinder and Cone; Draw the sectional orthographic views of geometrical solids, objects from industry and dwellings (foundation to slab only)

Sections of Solids:

Introduction

Sections and sectional views are used to show hidden detail more clearly. They are created by using a cutting plane to cut the object.

A section is a view of no thickness and shows the outline of the object at the cutting plane. Visible outlines beyond the cutting plane are not drawn.

A sectional view, displays the outline of the cutting plane and all visible outlines which can be seen beyond the cutting plane.

Improve visualization of interior features. Section views are used when important hidden details are in the interior of an object. These details appear as hidden lines in one of the orthographic principal views; therefore, their shapes are not very well described by pure orthographic projection.

Types of Section Views

- Full sections
- Half sections
- Offset sections
- Revolved sections
- Removed sections
- Broken-out sections



Cutting Plane

- Section views show how an object would look if a cutting plane (or saw) cut through the object and the material in front of the cutting plane was discarded

Representation of cutting plane

According to drawing standards cutting plane is represented by chain line with alternate long dash and dot. The two ends of the line should be thick.

Full Section View

- In a full section view, the cutting plane cuts across the entire object
- Note that hidden lines become visible in a section view

Hatching

On sections and sectional views solid area should be hatched to indicate this fact. Hatching is drawn with a thin continuous line, equally spaced (preferably about 4 mm apart, though never less than 1 mm) and preferably at an angle of 45°.

(i) Hatching a single object: When you are hatching an object, but the objects have areas that are separated. All areas of the object should be hatched in the same direction and with the same spacing.

(ii) Hatching Adjacent objects: When hatching assembled parts, the direction of the hatching should ideally be reversed on adjacent parts. If more than two parts are adjacent, then the hatching should be staggered to emphasize the fact that these parts are separate.

Problem: A square prism of base side on 30 mm and axis length 60 mm is resting on HP on one of its bases, with a base side inclined at 30° to VP. It is cut by a plane inclined at 40° to HP and perpendicular to VP and is bisecting the axis of the prism. Draw its front view, sectional top view and true shape of section.

Solution: Draw the projections of the prism in the given position. The top view is drawn and the front view is projected.

To draw the cutting plane, front view and sectional top view

1. Draw the Vertical Trace (VT) of the cutting plane inclined at 40° to XY line and passing through the midpoint of the axis.

2. As a result of cutting, longer edge $a' p'$ is cut, the end a' has been removed and the new corner l' is obtained.

3. Similarly $2'$ is obtained on longer edge $b' q'$, $3'$ on $c' r'$ and $4'$ on $d' s'$,

4. Show the remaining portion in front view by drawing dark lines.

5. Project the new points $1', 2', 3'$ and $4'$ to get $1, 2, 3$ and 4 in the top view of the prism, which are coinciding with the bottom end of the longer edges p, q, r and s respectively.

6. Show the sectional top view or apparent section by joining $1, 2, 3$ and 4 by drawing hatching lines.

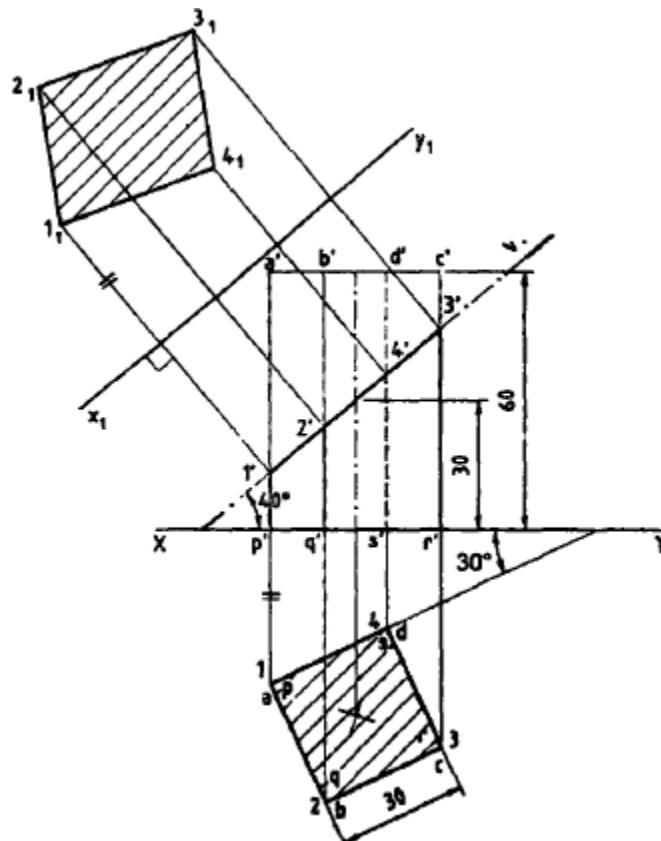


Fig. 4.1

To draw the true shape of a section

1. Consider an auxiliary inclined plane parallel to the cutting plane and draw the new reference line $X_1 Y_1$ parallel to VT of the cutting plane at an arbitrary distance from it.

2. Draw projectors passing through $1', 2', 3'$ and $4'$ perpendicular to $X_1 Y_1$ line.

3. The distance of point 1 in top view from XY line is measured and marked from $X_1 Y_1$ in the projector passing through $1'$ to get $1_1'$. This is repeated to get the other points $2_1, 3_1$ and 4_1 .

4. Join these points to get the true shape of section as shown by drawing the hatching lines.

Development of surfaces of various solids:

Introduction

A layout of the complete surface of a three dimensional object on a plane is called the development of the surface or flat pattern of the object. The development of surfaces is very important in the fabrication of articles made of sheet metal.

The objects such as containers, boxes, boilers, hoppers, vessels, funnels, trays etc., are made of sheet metal by using the principle of development of surfaces.

In making the development of a surface, an opening of the surface should be determined first. Every line used in making the development must represent the true length of the line (edge) on the object.

The steps to be followed for making objects, using sheet metal are given below:

1. Draw the orthographic views of the object to full size.
2. Draw the development on a sheet of paper.
3. Transfer the development to the sheet metal.
4. Cut the development from the sheet.
5. Form the shape of the object by bending.
6. Join the closing edges.

Methods of Development

The method to be followed for making the development of a solid depends upon the nature of its lateral surfaces. Based on the classification of solids, the following are the methods of development.

1. Parallel-line Development

It is used for developing prisms and single curved surfaces like cylinders in which all the edges / generators of lateral surfaces are parallel to each other.

2. Radial-line Development

It is employed for pyramids and single curved surfaces like cones in which the apex is taken as centre and the slant edge or generator (which are the true lengths) as radius for its development.

Development of Prism

To draw the development of a square prism of side of base 30 mm and height 50 mm.

Construction:

1. Assume the prism is resting on its base on H.P. with an edge of the base parallel to V.P and draw the orthographic views of the square prism.
2. Draw the stretch-out line 1-1 (equal in length to the circumference of the square prism) and mark off the sides of the base along this line in succession i.e. 1-2, 2-3, 3-4 and 4-1.
3. Erect perpendiculars through 1, 2, 3 etc., and mark the edges (folding lines) I-A, 2-B, etc., equal to the height of the prism 50 mm.
4. Add the bottom and top bases 1234 and ABCD by the side of an)' of the base edges.

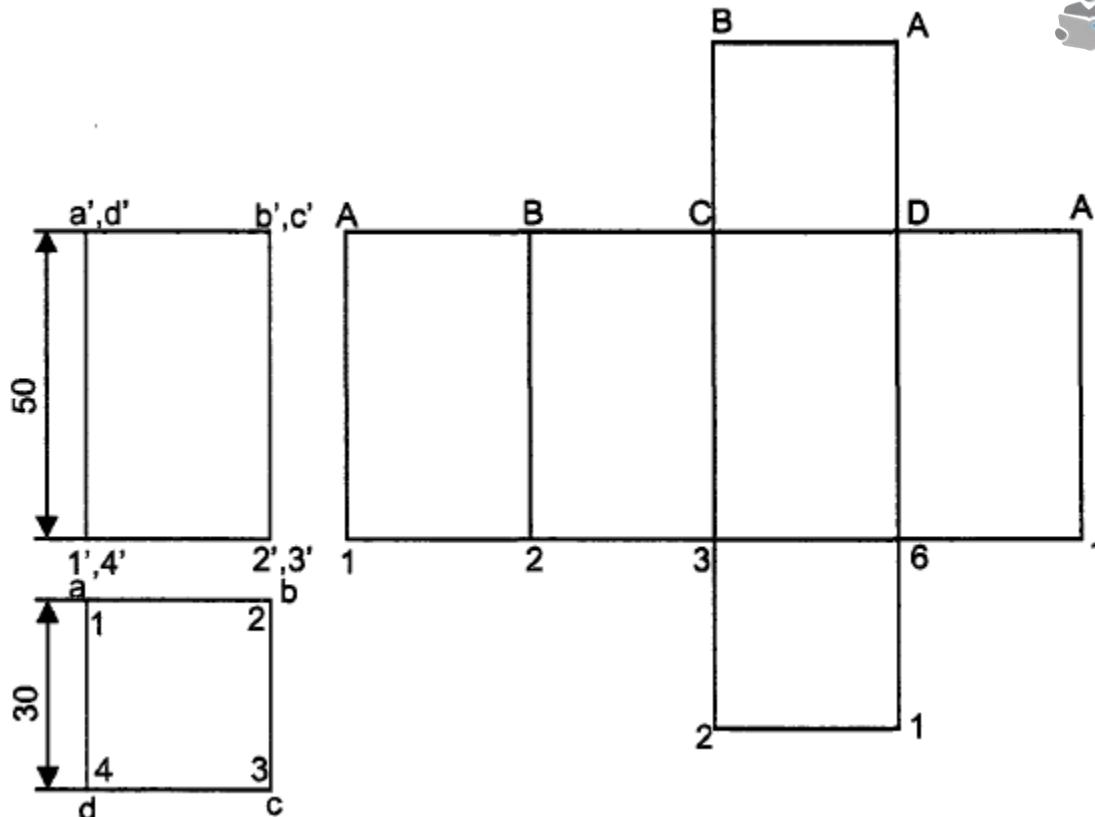


Fig. 4.2 Development of Prism

Development of a Cylinder

Construction:

Figure shows the development of a cylinder. In this the length of the rectangle representing the development of the lateral surface of the cylinder is equal to the circumference (πd here d is the diameter of the cylinder) of the circular base.

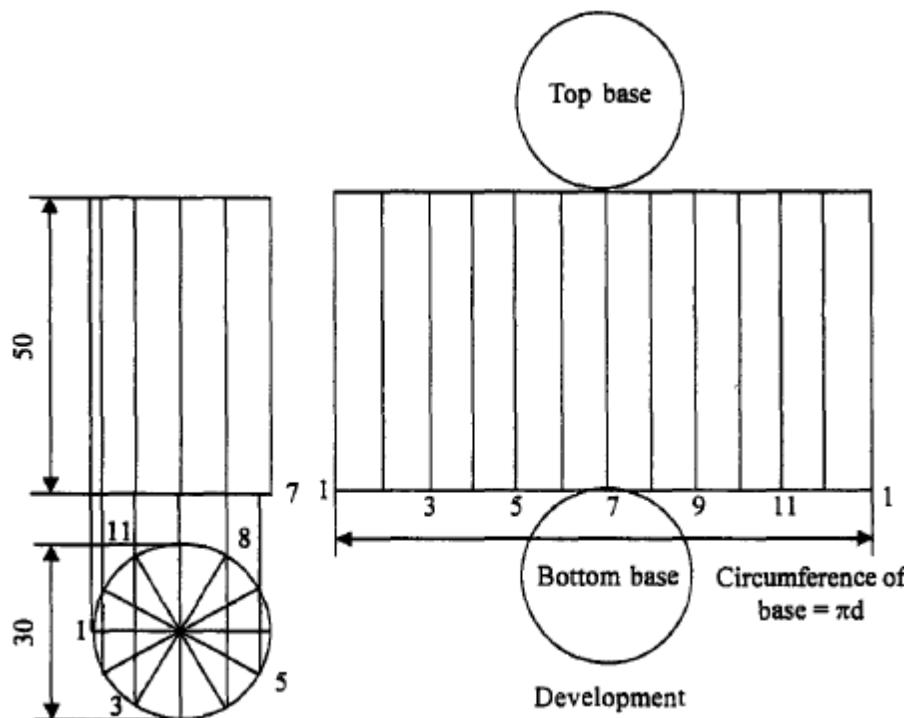


Fig. 4.3 Development of Cylinder

Development of a Pyramid

Problem: Development of a square pyramid with side of base 30 mm and height 60 mm.

Construction:

1. Draw the views of the pyramid assuming that it is resting on H.P and with an edge of the base parallel to V.P.
2. Determine the true length oa of the slant edge.

Note:

In the orientation given for the solid, all the slant edges are inclined to both H.P and V.P. Hence, neither the front view nor the top view provides the true length of the slant edge. To determine the true length of the slant edge, say OA, rotate oa till it is parallel to XY to the position oa_1' . Through a_1' , draw a projector to meet the line XY at a_1 . Then $o_1'a_1'$ all represents the true length of the slant edge OA. This method of determining the true length is also known as rotation method.

3. with centre O and radius $o_1'a_1'$ draw an arc.
4. Starting from A along the arc, mark the edges of the base i.e. AB, BC, CD and DA.
5. Join O to A, B, C, etc., representing the lines of folding and thus completing the development.

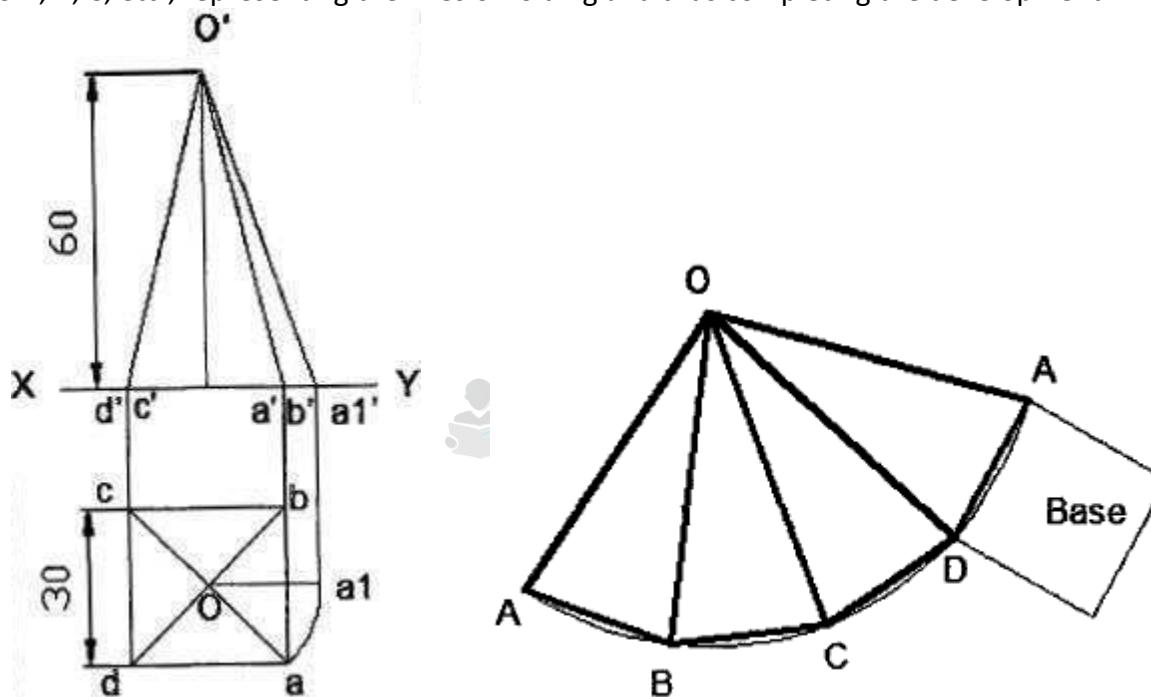


Fig. 4.4 Development of Square Pyramid

Development of Pentagonal Pyramid

Construction:

1. Draw the orthographic views of the pyramid ABCDE with its base on H.P and axis parallel to V.P.
2. With centre O of the pyramid and radius equal to the true length of the slant edge draw an arc.
3. Mark off the edges starting from A along the arc and join them to O representing the lines of folding.
4. Add the base at a suitable location.

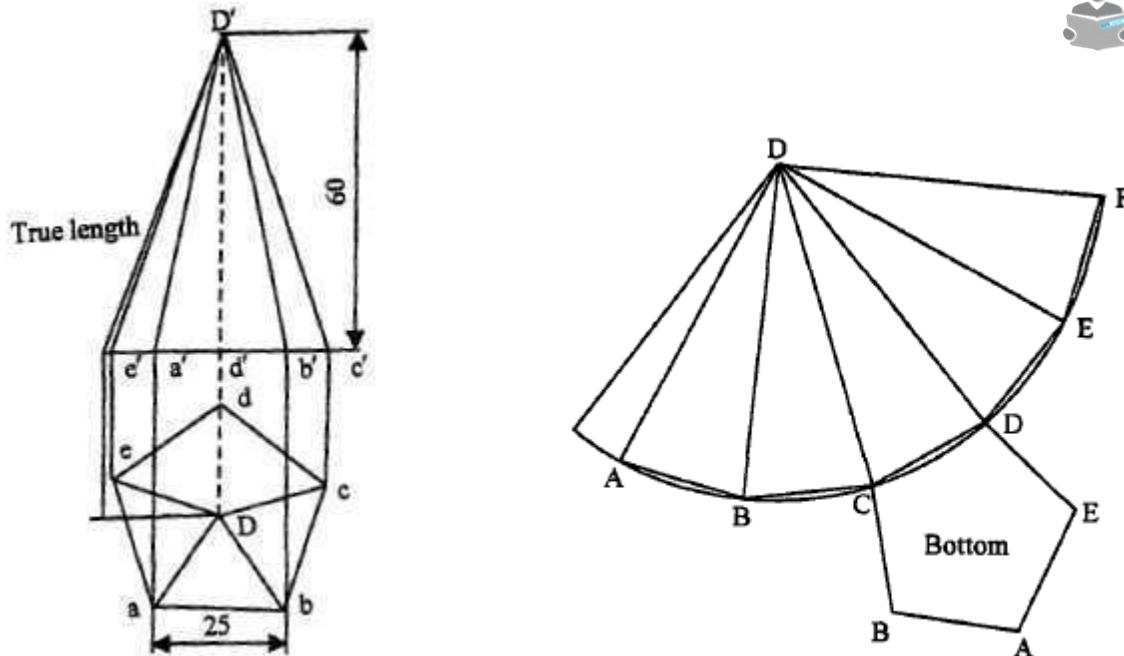


Fig. 4.5 Development of Pentagonal Pyramid

Development of a Cone

Construction:

The development of the lateral surface of a cone is a sector of a circle. The radius and length of the arc are equal to the slant height and circumference of the base of the cone respectively. The included angle of the sector is given by $(r / s) \times 360^\circ$, where r is the radius of the base of the cone and s is the true length.

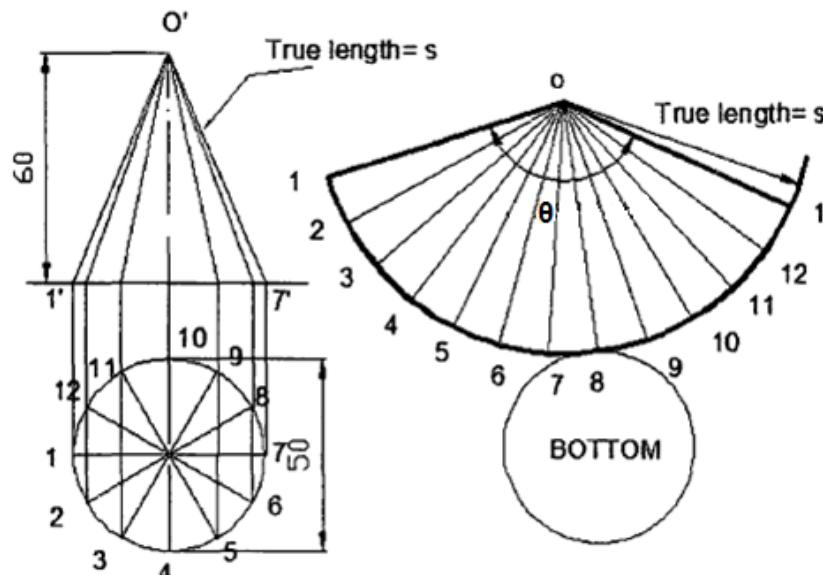


Fig. 4.6 Development of Cone

$$\theta = (r / s) \times 360^\circ$$

Where, r = Radius of base of Cone

s = Slant Height



RGPVNOTES.IN

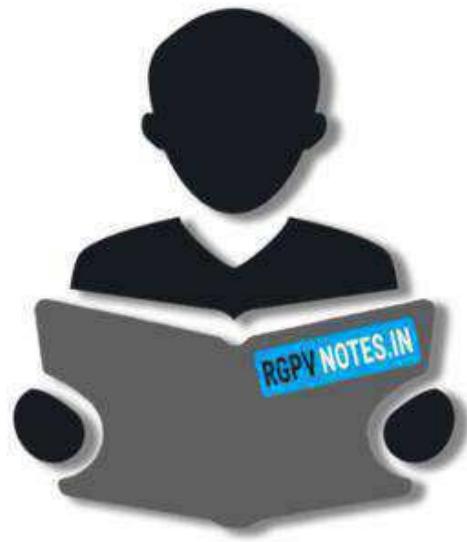
We hope you find these notes useful.

You can get previous year question papers at
<https://qp.rgpvnotes.in> .

If you have any queries or you want to submit your
study notes please write us at
rgpvnotes.in@gmail.com



LIKE & FOLLOW US ON FACEBOOK
facebook.com/rgpvnotes.in



RGPVNOTES.IN

Program : **B.Tech**

Subject Name: **Basic Electrical & Electronics Engineering**

Subject Code: **BT-104**

Semester: **1st**



LIKE & FOLLOW US ON FACEBOOK
facebook.com/rgpvnotes.in

NUMBER SYSTEMS

NUMBER SYSTEM :- A number system is a code having an assigned symbol for each distinct magnitude. The symbols are called “digits”. The number of digits in a number system will determine the base of the system. In all number systems, the weight of a number depends on its relative position.

BASE or RADIX : - The base or radix of a number system is the total number of different digits or basic symbols used in a number system. In the binary system we have 0 & 1 as digits , so the base or radix is 2. In the decimal system we have 10 digits ie. 0 through 9, so the base or radix is 10.

BINARY SYSTEM : - This number system has a base or radix of 2. The symbols or digits used in this system are 0 & 1.

OCTAL SYSTEM : - This number system has a base or radix of 8. The symbols or digits used in this system are 0 through 7. (0, 1, 2, 3, 4, 5, 6, 7)

DECIMAL SYSTEM : - This number system has a base or radix of 10. The symbols or digits used in this system are 0 through 9. (0, 1, 2, 3, 4, 5, 6, 7, 8, 9)

HEXA DECIMAL SYSTEM : - This number system has a base or radix of 16. The symbols or digits used in this system are 0 through F. (0, 1, 2, 3, 4, 5, 6, 7, 8, 9, A, B, C, D, E, F)

CODE CONVERSIONS

(1) **BINARY TO DECIMAL CONVERSION:-**

$$(i) (101111.11010)_2 = (?)_{10}$$

Integral part :
$$\begin{aligned} (101111)_2 &= (1 \times 2^5) + (0 \times 2^4) + (1 \times 2^3) + (1 \times 2^2) + (1 \times 2^1) + (1 \times 2^0) \\ &= 32 + 0 + 8 + 4 + 2 + 1 \\ &= 47 \end{aligned}$$

ie. $(101111)_2 = (47)_{10}$

Fractional Part :
$$\begin{aligned} (0.11010)_2 &= (1 \times 1/2) + (1 \times 1/4) + (0 \times 1/8) + (1 \times 1/16) + (0 \times 1/32) \\ &= 0.5 + 0.25 + 0 + 0.0625 + 0 \\ &= 0.8125 \end{aligned}$$

ie. $(0.11010)_2 = (0.8125)_{10}$

Thus $(101111.11010)_2 = (47.8125)_{10}$

$$(ii) (111010.101101)_2 = (?)_{10}$$

Integral part :
$$\begin{aligned} (111010)_2 &= (1 \times 2^5) + (1 \times 2^4) + (1 \times 2^3) + (0 \times 2^2) + (1 \times 2^1) + (0 \times 2^0) \\ &= 32 + 16 + 8 + 0 + 2 + 0 \\ &= 58 \end{aligned}$$

ie. $(111010)_2 = (58)_{10}$

Fractional Part :
$$\begin{aligned} (0.101101)_2 &= (1 \times 1/2) + (0 \times 1/4) + (1 \times 1/8) + (1 \times 1/16) + (0 \times 1/32) + (0 \times 1/64) \\ &= 0.5 + 0 + 0.125 + 0.0625 + 0 + 0.015625 \\ &= 0.703125 \end{aligned}$$

ie. $(0.101101)_2 = (0.703125)_{10}$

Thus $(111010.101101)_2 = (58.703125)_{10}$

(2) OCTAL TO DECIMAL CONVERSION:-

$$(i) (457.245)_8 = (?)_{10}$$

Integral part :

$$\begin{aligned}(457)_8 &= (4 \times 8^2) + (5 \times 8^1) + (7 \times 8^0) \\ &= 256 + 40 + 7 \\ &= 303\end{aligned}$$

$$\text{ie. } (457)_8 = (303)_{10}$$

Fractional Part :

$$\begin{aligned}(0.245)_8 &= (2 \times 1/8) + (4 \times 1/64) + (5 \times 1/512) \\ &= 0.25 + 0.0625 + 0.0097656 \\ &= 0.3222656\end{aligned}$$

$$\text{ie. } (0.245)_8 = (0.3222656)_{10}$$

$$\text{Thus } (457.245)_8 = (303.3222656)_{10}$$

$$(ii) (1427.3426)_8 = (?)_{10}$$

Integral part :

$$\begin{aligned}(1427)_8 &= (1 \times 8^3) + (4 \times 8^2) + (2 \times 8^1) + (7 \times 8^0) \\ &= 512 + 256 + 16 + 7 \\ &= 791\end{aligned}$$

$$\text{ie. } (1427)_8 = (791)_{10}$$

Fractional Part :

$$\begin{aligned}(0.3426)_8 &= (3 \times 1/8) + (4 \times 1/64) + (2 \times 1/512) + (6 \times 1/4096) \\ &= 0.375 + 0.0625 + 0.00391 + 0.001465 \\ &= 0.442875\end{aligned}$$

$$\text{ie. } (0.3426)_8 = (0.442875)_{10}$$

$$\text{Thus } (1427.3426)_8 = (791.442875)_{10}$$

(3) HEXADECIMAL TO DECIMAL CONVERSION :-

$$(i) (F9AC.5D8B)_{16} = (?)_{10}$$

Integral part :

$$\begin{aligned}(F9AC)_{16} &= (F \times 16^3) + (9 \times 16^2) + (A \times 16^1) + (C \times 16^0) \\ &= (15 \times 16^3) + (9 \times 16^2) + (10 \times 16^1) + (12 \times 16^0) \\ &= 61440 + 2304 + 160 + 12 \\ &= 63916\end{aligned}$$

$$\text{ie. } (F9AC)_{16} = (63916)_{10}$$

Fractional Part :

$$\begin{aligned}(0.5D8B)_{16} &= (5 \times 1/16) + (D \times 1/16^2) + (8 \times 1/16^3) + (B \times 1/16^4) \\ &= 0.3125 + 0.051 + 0.00195 + 0.000168 \\ &= 0.36562\end{aligned}$$

$$\text{ie. } (0.5D8B)_{16} = (0.36562)_{10}$$

$$\text{Thus } (F9AC.5D8B)_{16} = (63916.36562)_{10}$$

$$(ii) (9EA6.2FA)_{16} = (?)_{10}$$

Integral part :

$$\begin{aligned}(9EA6)_{16} &= (9 \times 16^3) + (E \times 16^2) + (A \times 16^1) + (6 \times 16^0) \\ &= (9 \times 16^3) + (14 \times 16^2) + (10 \times 16^1) + (6 \times 1) \\ &= 36864 + 3584 + 160 + 6 \\ &= 40614\end{aligned}$$

$$\text{ie. } (9EA6)_{16} = (40614)_{10}$$

Fractional Part :

$$\begin{aligned}
 (0.2FA)_{16} &= (2 \times 1/16) + (F \times 1/16^2) + (A \times 1/16^3) \\
 &= (2 \times 1/16) + (15 \times 1/256) + (10 \times 1/4096) \\
 &= 0.125 + 0.0586 + 0.00244 \\
 &= 0.18604
 \end{aligned}$$

ie. $(0.2FA)_{16} = (0.18604)_{10}$

Thus $(9EA6.2FA)_8 = (40614.18604)_{10}$

(1) DECIMAL TO BINARY CONVERSION :-

(i) $(47.8125)_{10} = (?)_2$

Integral part :

2	47	1
2	23	1
2	11	1
2	5	1
2	2	0
	1	

ie. $(47)_{10} = (101111)_2$

Fractional Part :

$$\begin{array}{lcl}
 (0.8125 \times 2) = 1.625 & \rightarrow 1 & \\
 (0.625 \times 2) = 1.25 & \rightarrow 1 & \\
 (0.25 \times 2) = 0.5 & \rightarrow 0 & \\
 (0.5 \times 2) = 1.0 & \rightarrow 1 &
 \end{array}$$

ie. $(0.8125)_{10} = (0.1101)_2$

Thus $(47.8125)_{10} = (101111.1101)_2$

(ii) $(58.703125)_{10} = (?)_2$

Integral part :

2	58	0
2	29	1
2	14	0
2	7	1
2	3	1
	1	

ie. $(58)_{10} = (111010)_2$

Fractional Part :

$$\begin{array}{lcl}
 (0.703125 \times 2) = 1.40625 & \rightarrow 1 & \\
 (0.40625 \times 2) = 0.8125 & \rightarrow 0 & \\
 (0.8125 \times 2) = 1.625 & \rightarrow 1 & \\
 (0.625 \times 2) = 1.25 & \rightarrow 1 & \\
 (0.25 \times 2) = 0.5 & \rightarrow 0 & \\
 (0.5 \times 2) = 1.0 & \rightarrow 1 &
 \end{array}$$

ie. $(0.703125)_{10} = (0.101101)_2$

Thus $(58.703125)_{10} = (111010.101101)_2$

(2) DECIMAL TO OCTAL CONVERSION :-

$$(i) (303.3222656)_{10} = (?)_8$$

Integral part :

8	303	7	↑
8	37	5	
4			

$$\text{ie. } (303)_{10} = (457)_8$$

Fractional Part :

(0.3222656×8)	=	2.5781248	$\rightarrow 2$
(0.5781248×8)	=	4.6249984	$\rightarrow 4$
(0.6249984×8)	=	4.9999872	$\rightarrow 4$
(0.9999872×8)	=	7.9998976	$\rightarrow 7$
(0.9998976×8)	=	7.9991808	$\rightarrow 7$
(0.9991808×8)	=	7.9934464	$\rightarrow 7$

$$\text{ie. } (0.3222656)_{10} = (0.244777)_8$$

$$\text{Thus } (303.3222656)_{10} = (457.244777)_8$$

$$(ii) (791.442875)_{10} = (?)_8$$

Integral part :

8	791	7	↑
8	98	2	
8	12	4	
1			

$$\text{ie. } (791)_{10} = (1427)_8$$

Fractional Part :

(0.442875×8)	=	3.543	$\rightarrow 3$
(0.543×8)	=	4.344	$\rightarrow 4$
(0.344×8)	=	2.752	$\rightarrow 2$
(0.752×8)	=	6.016	$\rightarrow 6$
(0.016×8)	=	0.128	$\rightarrow 0$
(0.128×8)	=	1.024	$\rightarrow 1$

$$\text{ie. } (0.442875)_{10} = (0.342601)_8$$

$$\text{Thus } (791.442875)_{10} = (1427.342601)_8$$

(3) DECIMAL TO HEXADECIMAL CONVERSION :-

$$(i) (63916.36562)_{10} = (?)_{16}$$

Integral part :

16	63916	12	$\longrightarrow C$	↑
16	3994	10	$\longrightarrow A$	
16	249	9	$\longrightarrow 9$	
	15		$\longrightarrow F$	

$$\text{ie. } (63916)_{10} = (F9AC)_{16}$$

Fractional Part :

$$\begin{aligned}
 (0.36562 \times 16) &= 5.84992 \rightarrow 5 \rightarrow 5 \\
 (0.84992 \times 16) &= 13.59872 \rightarrow 13 \rightarrow D \\
 (0.59872 \times 16) &= 9.57952 \rightarrow 9 \rightarrow 9 \\
 (0.57952 \times 16) &= 9.27232 \rightarrow 9 \rightarrow 9 \\
 (0.27232 \times 16) &= 4.35712 \rightarrow 4 \rightarrow 4
 \end{aligned}$$

ie. $(0.36562)_{10} = (0.5D994)_{16}$

Thus $(63916.36562)_{10} = (F9AC.5D994)_{16}$

(ii) $(40614.18604)_{10} = (?)_{16}$

Integral part :

16	40614	6	→ 6
16	2538	10	→ A
16	158	14	→ E
	9		→ 9

ie. $(40614)_{10} = (9EA6)_{16}$

Fractional Part :

$$\begin{aligned}
 (0.18604 \times 16) &= 2.97664 \rightarrow 2 \rightarrow 2 \\
 (0.97664 \times 16) &= 15.62624 \rightarrow 15 \rightarrow F \\
 (0.62624 \times 16) &= 10.01984 \rightarrow 10 \rightarrow A \\
 (0.01984 \times 16) &= 0.31744 \rightarrow 0 \rightarrow 0 \\
 (0.31744 \times 16) &= 5.07904 \rightarrow 5 \rightarrow 5
 \end{aligned}$$

ie. $(0.18604)_{10} = (0.2FA05)_{16}$

Thus $(40614.18604)_{10} = (9EA6.2FA05)_{16}$

**(1) OCTAL TO BINARY CONVERSION :-**

(i) $(457.245)_8 = (?)_2$

Integral part :

$$\begin{aligned}
 (457)_8 &= \overbrace{4}^{\{100\}} \overbrace{5}^{\{101\}} \overbrace{7}^{\{111\}} \\
 &= (100101111)_2
 \end{aligned}$$

ie. $(457)_8 = (100101111)_2$

Fractional Part :

$$\begin{aligned}
 (0.245)_8 &= \overbrace{2}^{\{010\}} \overbrace{4}^{\{100\}} \overbrace{5}^{\{101\}} \\
 &= (0.010100101)_2
 \end{aligned}$$

ie. $(0.245)_8 = (0.010100101)_2$

Thus $(457.245)_8 = (100101111.010100101)_2$

(ii) $(1427.3426)_8 = (?)_2$

Integral part :

$$\begin{aligned}
 (1427)_8 &= \overbrace{1}^{\{001\}} \overbrace{4}^{\{100\}} \overbrace{2}^{\{010\}} \overbrace{7}^{\{111\}} \\
 &= (1100010111)_2
 \end{aligned}$$

ie. $(1427)_8 = (1100010111)_2$

Fractional Part :

$$(0.3426)_8 = \frac{3}{011} \quad \frac{4}{100} \quad \frac{2}{010} \quad \frac{6}{110} \\ = (0.01110001011)_2$$

ie. $(0.3426)_8 = (0.01110001011)_2$

Thus $(1427.3426)_8 = (1100010111.01110001011)_2$

(2) HEXADECIMAL TO BINARY CONVERSION :-

(i) $(F9AC.5D8B)_{16} = (?)_2$

Integral part :

$$(F9AC)_{16} = \frac{F}{1111} \quad \frac{9}{1001} \quad \frac{A}{1010} \quad \frac{C}{1100} = (111100110101100)_2$$

ie. $(F9AC)_{16} = (111100110101100)_2$

Fractional Part :

$$(0.5D8B)_{16} = \frac{5}{0101} \quad \frac{D}{1101} \quad \frac{8}{1000} \quad \frac{B}{1011} = (0.0101110110001011)_2$$

Thus $(F9AC.5D8B)_{16} = (111100110101100.0101110110001011)_2$

(ii) $(9EA6.2FA)_{16} = (?)_2$

Integral part :



$$(9EA6)_{16} = \frac{9}{1001} \quad \frac{E}{1110} \quad \frac{A}{1010} \quad \frac{6}{0110} = (1001111010100110)_2$$

ie. $(9EA6)_{16} = (1001111010100110)_2$

Fractional Part :

$$(0.2FA)_{16} = \frac{2}{0010} \quad \frac{F}{1111} \quad \frac{A}{1010} = (0.00101111010)_2$$

Thus $(9EA6.2FA)_{16} = (1001111010100110.00101111010)_2$

(1) BINARY TO OCTAL CONVERSION :-

(i) $(100101111.010100101)_2 = (?)_8$

Integral part :

$$\overleftarrow{(100, 101, 111)}_2 = \{\frac{100}{4}, \frac{101}{5}, \frac{111}{7}\} = (457)_8$$

ie. $(100101111)_2 = (457)_8$

Fractional Part :-

$$\overrightarrow{(0.010, 100, 101)}_2 = \{\frac{010}{2}, \frac{100}{4}, \frac{101}{5}\} = (245)_8$$

ie. $(100101111)_2 = (0.245)_8$

Thus $(100101111.010100101)_2 = (457.245)_8$

$$(ii) (1100010111 \cdot 01110001011)_2 = (?)_8$$

Integral part :

$$\overleftarrow{(1, 100, 010, 111)}_2 = \{ \underbrace{001}_1, \underbrace{100}_4, \underbrace{010}_2, \underbrace{111}_7 \} = (1427)_8$$

$$\text{ie. } (1100010111)_2 = (1427)_8$$

Fractional Part :

$$\overrightarrow{(0.011, 100, 010, 11)}_2 = \{ 011 \underbrace{}_3, \underbrace{100}_4, \underbrace{010}_2, \underbrace{110}_6 \} = (0.3426)_8$$

$$\text{ie. } (0.01110001011)_2 = (0.3426)_8$$

$$\text{Thus } (1100010111 \cdot 01110001011)_2 = (1427 \cdot 3426)_8$$

(2) BINARY TO HEXADECIMAL CONVERSION :-

$$(i) (1001111010100110 \cdot 001011111010)_2 = (?)_{16}$$

Integral part :

$$\overleftarrow{(1001111010100110)}_2 = \{ 1001, 1110, 1010, 0110 \}$$

$$= (\underbrace{1001}_9, \underbrace{1110}_E, \underbrace{1010}_A, \underbrace{0110}_6)_2$$

$$\text{ie. } (1001111010100110)_2 = (9EA6)_{16}$$

Fractional Part :

$$\overrightarrow{(0.00101111010)}_2 = \{ 0010 \underbrace{}_2, 1111 \underbrace{}_F, 1010 \underbrace{}_A \}$$

$$\text{ie. } (0.00101111010)_2 = (0.2FA)_{16}$$

$$\text{Thus } (1001111010100110 \cdot 001011111010)_2 = (9EA6 \cdot 2FA)_{16}$$

$$(ii) (1111100110101100 \cdot 010111011000101100)_2 = (?)_{16}$$

Integral part :

$$\overleftarrow{(1111100110101100)}_2 = \{ 1111, 1001, 1010, 1100 \}$$

$$= (\underbrace{1111}_F, \underbrace{1001}_9, \underbrace{1010}_A, \underbrace{1100}_C)_2$$

$$\text{ie. } (1111100110101100)_2 = (F9AC)_{16}$$

Fractional Part :

$$\overrightarrow{(0.010111011000101100)}_2 = \{ 0101 \underbrace{}_5, 1101 \underbrace{}_D, 1000 \underbrace{}_8, 1011 \underbrace{}_B, 0000 \underbrace{}_0 \}$$

$$\text{ie. } (0.010111011000101100)_2 = (0.5D8B)_{16}$$

$$\text{Thus } (1111100110101100 \cdot 0101110110001011)_{16} = (F9AC \cdot 5D8B)_{16}$$

(1) HEXADECIMAL TO OCTAL CONVERSION :-

(i) ($F9AC . 5D8B$)₁₆ = (?)₈

Integral part :

$$(F9AC)_{16} = \{ \overline{\text{F}} \quad \overline{\text{9}} \quad \overline{\text{A}} \quad \overline{\text{C}} \} = (1111100110101100)_2$$

$$\begin{aligned} \text{ie. } (F9AC)_{16} &= (\overleftarrow{1, 111, 100, 110, 101, 100})_2 \\ &= (\overbrace{001, \overbrace{111, \overbrace{100, \overbrace{110, \overbrace{101, \overbrace{100}}}}}}{1, 7, 4, 6, 5, 4})_2 = 174654 \end{aligned}$$

$$\text{ie. } (F9AC)_{16} = (174654)_8$$

Fractional Part :

$$\begin{aligned} (0 . 5D8B)_{16} &= \overline{\{ \overline{5} \quad \overline{D} \quad \overline{8} \quad \overline{B} \}} = (\overrightarrow{010, 111, 011, 000, 101, 100})_2 \\ &= (\overbrace{010, \overbrace{111, \overbrace{011, \overbrace{000, \overbrace{101, \overbrace{100}}}}}}{2, 7, 3, 0, 5, 4})_2 \\ &= (0 . 273054)_8 \end{aligned}$$

$$\text{ie. } (0.5D8B) = (0 . 273054)_8$$

$$\text{Thus } (F9AC . 5D8B)_{16} = (174654 . 273054)_8$$

(ii) ($9EA6 . 2FA$)₁₆ = (?)₈

Integral part :

$$\begin{aligned} (9EA6)_{16} &= \overline{\{ \overline{9} \quad \overline{E} \quad \overline{A} \quad \overline{6} \}} \\ &= (1001111010100110)_2 \end{aligned}$$

$$\begin{aligned} \text{ie. } (9EA6)_{16} &= (\overleftarrow{1, 001, 111, 010, 100, 110})_2 \\ &= (\overbrace{1, \overbrace{001, \overbrace{111, \overbrace{010, \overbrace{100, \overbrace{110}}}}}}{1, 1, 7, 2, 4, 6})_2 \\ &= (117246)_8 \end{aligned}$$

$$\text{ie. } (9EA6)_{16} = (117246)_8$$

Fractional Part :

$$\begin{aligned} (0 . 2FA)_{16} &= \overline{\{ \overline{2} \quad \overline{F} \quad \overline{A} \}} \\ &= (0 . 001011111010)_2 \\ &= (0 . \overline{001, 011, 111, 010})_2 \\ &= (\overbrace{0, \overbrace{001, \overbrace{011, \overbrace{111, \overbrace{010}}}}}}{1, 3, 7, 2})_2 \\ &= (0 . 1372)_8 \end{aligned}$$

$$\text{ie. } (0 . 2FA)_{16} = (0 . 1372)_8$$

$$\text{Thus } (9EA6 . 2FA)_{16} = (117246 . 1372)_8$$

(2) OCTAL TO HEXADECIMAL CONVERSION :-

$$(i) (174654.273054)_8 = (?)_{16}$$

Integral part :

$$\begin{aligned} (174654)_8 &= \left\{ \begin{array}{c} 1 \\ 001 \\ \hline 7 \\ 111 \\ \hline 4 \\ 100 \\ \hline 6 \\ 110 \\ \hline 5 \\ 101 \\ \hline 4 \\ 100 \end{array} \right\} \\ &= (\underbrace{0}_{0}, \underbrace{111}_{F}, \underbrace{1001}_{9}, \underbrace{1010}_{A}, \underbrace{1100}_{C})_2 \end{aligned}$$

$$\text{ie. } (174654)_8 = (F9AC)_{16}$$

Fractional Part :

$$\begin{aligned} (0.273054)_8 &= \left\{ \begin{array}{c} 2 \\ 010 \\ \hline 7 \\ 111 \\ \hline 3 \\ 011 \\ \hline 0 \\ 000 \\ \hline 5 \\ 101 \\ \hline 4 \\ 100 \end{array} \right\} \\ &= (0. \overbrace{0101, 1101, 1000, 1011, 0000}^{\rightarrow})_{16} \\ &\quad \overbrace{\hphantom{0.0101, 1101, 1000, 1011, 0000}}^{5D8B0} \overbrace{\hphantom{0.0101, 1101, 1000, 1011, 0000}}^{8} \overbrace{\hphantom{0.0101, 1101, 1000, 1011, 0000}}^{B} \overbrace{\hphantom{0.0101, 1101, 1000, 1011, 0000}}^{0} \end{aligned}$$

$$\text{ie. } (0.273054)_8 = (0.5D8B0)_{16}$$

$$\text{Thus } (174654.273054)_8 = (F9AC.5D8B0)_{16}$$

$$(ii) (117246.1372)_8 = (?)_{16}$$

Integral part :

$$\begin{aligned} (117246)_8 &= \left\{ \begin{array}{c} 1 \\ 001 \\ \hline 1 \\ 001 \\ \hline 7 \\ 111 \\ \hline 2 \\ 010 \\ \hline 4 \\ 100 \\ \hline 6 \\ 110 \end{array} \right\} \\ &= (\underbrace{0000, 1001, 1110, 1010, 0110}^{\leftarrow})_2 \\ &\quad \overbrace{\hphantom{0000, 1001, 1110, 1010, 0110}}^{0} \overbrace{\hphantom{0000, 1001, 1110, 1010, 0110}}^{9} \overbrace{\hphantom{0000, 1001, 1110, 1010, 0110}}^{E} \overbrace{\hphantom{0000, 1001, 1110, 1010, 0110}}^{A} \overbrace{\hphantom{0000, 1001, 1110, 1010, 0110}}^{6} \end{aligned}$$

$$\text{ie. } (117246)_8 = (09EA6)_{16}$$

Fractional Part :

$$\begin{aligned} (0.1372)_8 &= \left\{ \begin{array}{c} 1 \\ 001 \\ \hline 3 \\ 011 \\ \hline 7 \\ 111 \\ \hline 2 \\ 010 \end{array} \right\} \\ &= (0. \overbrace{0010, 1111, 1010, }^{\rightarrow})_{16} \\ &\quad \overbrace{\hphantom{0010, 1111, 1010, }}^{2} \overbrace{\hphantom{0010, 1111, 1010, }}^{F} \overbrace{\hphantom{0010, 1111, 1010, }}^{A} \end{aligned}$$

$$\text{ie. } (0.1372)_8 = (0.2FA)_{16}$$

$$\text{Thus } (117246.1372)_8 = (09EA6.2FA)_{16}$$

BINARY ARITHMETIC

(1) BINARY ADDITION :

(i) Add $(111011.1101)_2$ and $(11111.011)_2$

		1	1	1	0	1	1	.	1	1	0	1	1	Augend
+		0	1	1	1	1	1	.	0	1	1	0	0	Addend
	1	0	1	1	0	1	1	.	0	0	1	1	1	Sum

(ii) Add $(10110.1111)_2$ and $(1000111.1101)_2$

	0	0	1	0	1	1	0	.	1	1	1	1	1	Augend
+	1	0	0	0	1	1	1	.	1	1	0	1	1	Addend
	1	0	1	1	1	1	0	.	1	1	0	0	0	Sum

(2) BINARY SUBTRACTION :

(i) Subtract $(11111.011)_2$ from $(111011.1101)_2$

	1	1	1	0	1	1	.	1	1	0	1	1	Minuend
-	0	1	1	1	1	1	.	0	1	1	0	0	Subtrahend
	1	1	1	0	0	0	.	0	1	1	1	1	Difference

(ii) Subtract $(10110.1111)_2$ from $(1000111.1100)_2$

	1	0	0	0	1	1	1	.	1	1	0	0	0	Minuend
-	0	0	1	0	1	1	0	.	1	1	1	1	1	Subtrahend
	1	1	0	0	0	0	0	.	1	1	0	1	1	Difference

(3) BINARY MULTIPLICATION :

(i) Multiply $(1110.110)_2$ and $(1010.010)_2$

Multiplicand								Multiplier											
1	1	1	0	.	1	1	0	-	1	0	1	0	.	0	1	0			
										0	0	0	0	0	0	0	0		
									1	1	1	0	1	1	0				
									0	0	0	0	0	0	0				
									0	0	0	0	0	0	0				
									1	1	1	0	1	1	0				
									0	0	0	0	0	0	0				
									1	1	1	0	1	1	0				
									0	0	0	0	0	0	0				
									1	1	1	0	1	1	0				
									0	0	0	0	0	0	0				
									1	1	1	0	1	1	0				
									0	0	0	0	0	0	0				
									1	1	1	0	1	1	0				
									0	0	0	0	0	0	0				
									1	1	1	0	1	1	0				
									0	0	0	0	0	0	0				
									1	0	0	1	0	1	1	.	0	0	1
									0	0	0	0	0	0	0		1	1	0
									1	0	0	1	0	1	1	.	0	0	1
									0	0	0	0	0	0	0		1	1	0
									1	0	0	1	0	1	1	.	0	0	1
									0	0	0	0	0	0	0		1	1	0
									1	0	0	1	0	1	1	.	0	0	1
									0	0	0	0	0	0	0		1	1	0
									1	0	0	1	0	1	1	.	0	0	1
									0	0	0	0	0	0	0		1	1	0
									1	0	0	1	0	1	1	.	0	0	1
									0	0	0	0	0	0	0		1	1	0
									1	0	0	1	0	1	1	.	0	0	1
									0	0	0	0	0	0	0		1	1	0
									1	0	0	1	0	1	1	.	0	0	1
									0	0	0	0	0	0	0		1	1	0
									1	0	0	1	0	1	1	.	0	0	1
									0	0	0	0	0	0	0		1	1	0
									1	0	0	1	0	1	1	.	0	0	1
									0	0	0	0	0	0	0		1	1	0
									1	0	0	1	0	1	1	.	0	0	1
									0	0	0	0	0	0	0		1	1	0
									1	0	0	1	0	1	1	.	0	0	1
									0	0	0	0	0	0	0		1	1	0
									1	0	0	1	0	1	1	.	0	0	1
									0	0	0	0	0	0	0		1	1	0
									1	0	0	1	0	1	1	.	0	0	1
									0	0	0	0	0	0	0		1	1	0
									1	0	0	1	0	1	1	.	0	0	1
									0	0	0	0	0	0	0		1	1	0
									1	0	0	1	0	1	1	.	0	0	1
									0	0	0	0	0	0	0		1	1	0
									1	0	0	1	0	1	1	.	0	0	1
									0	0	0	0	0	0	0		1	1	0
									1	0	0	1	0	1	1	.	0	0	1
									0	0	0	0	0	0	0		1	1	0
									1	0	0	1	0	1	1	.	0	0	1
									0	0	0	0	0	0	0		1	1	0
									1	0	0	1	0	1	1	.	0	0	1
									0	0	0	0	0	0	0		1	1	0
									1	0	0	1	0	1	1	.	0	0	1
									0	0	0	0	0	0	0		1	1	0
									1	0	0	1	0	1	1	.	0	0	1
									0	0	0	0	0	0	0		1	1	0
									1	0	0	1	0	1	1	.	0	0	1
									0	0	0	0	0	0	0		1	1	0
									1	0	0	1	0	1	1	.	0	0	1
									0	0	0	0	0	0	0		1	1	0
									1	0	0	1	0	1	1	.	0	0	1
									0	0	0	0	0	0	0		1	1	0
									1	0	0	1	0	1	1	.	0	0	1
									0	0	0	0	0	0	0		1	1	0
									1	0	0	1	0	1	1	.	0	0	1
									0	0	0	0	0	0	0		1	1	0
									1	0	0	1	0	1	1	.	0	0	1
									0	0	0	0	0	0	0		1	1	0
									1	0	0	1	0	1	1	.	0	0	1
									0	0	0	0	0	0	0		1	1	0
									1	0	0	1	0	1	1	.	0	0	1
									0	0	0	0	0	0	0		1	1	0
									1	0									

(ii) Multiply $(1010.11)_2$ and $(111.01)_2$

Multiplicand							Multiplier								
1	0	1	0	.	1	1	1	1	1	1	.	0	1		
								1	0	1	0	1	1		
							0	0	0	0	0	0			
							1	0	1	0	1	1			Partial products
					1	0	1	0	1	1					
				1	0	1	0	1	1						
			1	0	1	0	1	1							
		1	0	0	1	1	0	1	.1	1	1	1			Final Product

(4) BINARY DIVISION:

(i) Divide $(110001)_2$ by $(111)_2$

$$\begin{array}{r}
 \text{Dividend} \\
 \text{Divisor } 111 \overline{)110001} \quad \text{Quotient} \\
 \underline{111} \\
 01010 \\
 \underline{111} \\
 000111 \\
 \underline{111} \\
 000 \quad \text{000 Remainder}
 \end{array}$$

(ii) Divide $(10100.110)_2$ by $(11.101)_2$

NOTE : { 10100.110 □ 11.101 } is the same as {10100110 □ 11101 }, therefore we can divide the numbers as shown below :

$$\begin{array}{r}
 \text{Dividend} \\
 \text{Divisor } 11101 \Big) 10100110 \quad \text{Quotient} \\
 \underline{11101} \\
 0110010 \\
 \underline{11101} \\
 000101010 \\
 \underline{11101} \\
 00000110100 \\
 \underline{11101} \\
 000000101110 \\
 \underline{11101} \\
 0000000100001 \quad \text{Remainder}
 \end{array}$$

BINARY SUBTRACTION USING COMPLEMENTARY NUMBERS

(1) BINARY SUBTRACTION USING 1's COMPLEMENT :

1's COMPLEMENT : The 1's complement of any binary number is obtained by subtracting every binary digit from 1 , for example , the 1's complement of the number 11011 is obtained as follows :
 $(11111 - 11011) = 00100$, therefore the 1's complement of $(11011)_2$ is $(00100)_2$.

The 1's complement is also obtained by complementing every digit of the given binary number, ie. The 1's complement of the number $(11111)_2$ is $(00000)_2$ & vice versa the 1's complement of $(00000)_2$ is $(11111)_2$. Negative numbers can be represented by 1's complement numbers, hence the process of subtraction in a processor can be carried out using an adder unit instead of a subtractor unit, as a result it minimizes the hardware in a computer.

(i) Subtract 25 from 45

$$\begin{array}{r}
 45 \\
 - 25 \\
 \hline
 20
 \end{array}
 \begin{array}{l}
 \xrightarrow{\quad} \text{binary equivalent is : } 101101 \xrightarrow{\quad} \text{write the positive number as it is} \\
 \xrightarrow{\quad} \text{binary equivalent is : } 011001 \xrightarrow{\quad} \text{write the 1's complement of the negative number}
 \end{array}$$

$$\begin{array}{r}
 \text{ie. } 45 \\
 - 25 \\
 \hline
 20
 \end{array}
 \begin{array}{l}
 \xrightarrow{\quad} \text{binary equivalent is : } \\
 \xrightarrow{\quad} \text{1's complement is : }
 \end{array}$$

	1	0	1	1	0	1
+	1	0	0	1	1	0
1	0	1	0	0	1	1

since carry is generated, it has to be taken as end-around carry as shown below :

	1	0	1	1	0	1
+	1	0	0	1	1	0
1	0	1	0	0	1	1
				1		1
	0	1	0	1	0	0

end-around carry

= 20

(ii) Subtract 15 from 31

$$\begin{array}{r}
 31 \\
 - 15 \\
 \hline
 16
 \end{array}
 \begin{array}{l}
 \xrightarrow{\quad} \text{binary equivalent is : } 11111 \xrightarrow{\quad} \text{write the positive number as it is} \\
 \xrightarrow{\quad} \text{binary equivalent is : } 01111 \xrightarrow{\quad} \text{write the 1's complement of the negative number}
 \end{array}$$

$$\begin{array}{r}
 \text{ie. } 31 \\
 - 15 \\
 \hline
 16
 \end{array}
 \begin{array}{l}
 \xrightarrow{\quad} \text{binary equivalent is } \\
 \xrightarrow{\quad} \text{1's complement is }
 \end{array}$$

	1	1	1	1	1
+	1	0	0	0	0
1	0	1	1	1	1

since carry is generated, it has to be taken as end-around carry as shown below :

	1	1	1	1	1
+	1	1	0	0	1
1	0	1	1	1	1
			1		1
	1	0	0	0	0

end-around carry

= 16

(iii) Subtract 31 from 15

$$\begin{array}{r}
 15 \\
 - 31 \\
 \hline
 - 16
 \end{array}
 \rightarrow \text{binary equivalent is : } 01111 \rightarrow \text{write the positive number as it is}$$

$$\begin{array}{r}
 \rightarrow \text{binary equivalent is : } 11111 \rightarrow \text{write the 1's complement of the negative number}
 \end{array}$$

$$\begin{array}{r}
 \text{ie. } 15 \\
 \begin{array}{r}
 \rightarrow \text{binary equivalent is} \\
 \begin{array}{r}
 - 31 \\
 \hline
 - 16
 \end{array}
 \end{array}
 \end{array}
 \rightarrow \text{1's complement is} \rightarrow$$

	0	1	1	1	1
+	0	0	0	0	0
	0	1	1	1	1

since carry is not generated, complement the result & attach a negative sign as shown below
 $\Rightarrow = (\square 10000) = (\square 16)$

(iv) Subtract 45 from 25

$$\begin{array}{r}
 25 \\
 - 45 \\
 \hline
 - 20
 \end{array}
 \rightarrow \text{binary equivalent is : } 011001 \rightarrow \text{write the positive number as it is}$$

$$\begin{array}{r}
 \rightarrow \text{binary equivalent is : } 101101 \rightarrow \text{write the 1's complement of the negative number}
 \end{array}$$

$$\begin{array}{r}
 \text{ie. } 25 \\
 \begin{array}{r}
 \rightarrow \text{binary equivalent is} \\
 \begin{array}{r}
 - 45 \\
 \hline
 - 20
 \end{array}
 \end{array}
 \end{array}
 \rightarrow \text{1's complement is} \rightarrow$$

	0	1	1	0	0	1
+	0	1	0	0	1	0
	1	0	1	0	1	1

since carry is not generated, complement the result & attach a negative sign as shown
 $\Rightarrow = (\square 010100) = (\square 20)$

(1) BINARY SUBTRACTION USING 2's COMPLEMENT :

2's COMPLEMENT : The 2's complement of any binary number is obtained by adding 1 to the 1's complement , for example , the 2's complement of the number 11011 is obtained as follows :
 The 1's complement of $(11011)_2$ is $(00100)_2$, the 2's complement is obtained by adding 1 to $(00100)_2$, ie.
 $(00100 + 1)_2 = (00101)_2$

The 2's complement is also obtained by writing the LSB of the given binary number as it is and complementing the rest of the digits . For example the 2's complement of $(11011)_2$ is $(00101)_2$. If the LSB is not a 1 but a 0 then all these initial 0's are retained unchanged & then the first 1 that is encountered is kept unchanged & the rest of the bits are complemented. For example the 2's complement of $(1101100)_2$ is $(0010100)_2$

Subtraction can be carried out through addition by using 2's complement numbers, hence subtraction in a processor can be carried out using an adder unit instead of a subtractor unit, as a result it minimizes the hardware in a computer. However the advantage of using 2's complement is that during the process of subtraction whenever a carry is generated , it need not be used as end-around carry but has to be just neglected. This means that the subtraction process using complementary numbers becomes simple. When a carry is not generated the resultant number will be a negative number.

(i) Subtract 25 from 45

$$\begin{array}{r}
 45 \\
 - 25 \\
 \hline
 20
 \end{array}
 \rightarrow \text{binary equivalent is : } 101101 \rightarrow \text{write the positive number as it is}$$

$$\begin{array}{r}
 \rightarrow \text{binary equivalent is : } 011001 \rightarrow \text{write the 2's complement of the negative number}
 \end{array}$$

$$\begin{array}{r}
 \text{ie. } 45 \\
 \begin{array}{r}
 \rightarrow \text{binary equivalent is} \\
 \begin{array}{r}
 - 25 \\
 \hline
 20
 \end{array}
 \end{array}
 \end{array}
 \rightarrow \text{2's complement is:} \rightarrow$$

	1	0	1	1	0	1
+	1	0	0	1	1	1
1	0	1	0	1	0	0

since carry is generated, it has to be neglected & result is taken as shown below
 $\text{ie. } (010100)_2 = (20)_{10}$

(ii) Subtract 15 from 31

$$\begin{array}{r}
 31 \\
 - 15 \\
 \hline
 16
 \end{array}
 \rightarrow \text{binary equivalent is : } 11111 \rightarrow \text{write the positive number as it is}$$

$$\rightarrow \text{binary equivalent is : } 01111 \rightarrow \text{write the 1's complement of the negative number}$$

$$\begin{array}{r}
 \text{ie. } 31 \\
 - 15 \\
 \hline
 16
 \end{array}
 \rightarrow \text{binary equivalent is} \rightarrow \boxed{\begin{array}{|c|c|c|c|c|c|} \hline & 1 & 1 & 1 & 1 & 1 \\ \hline + & 1 & 0 & 0 & 0 & 1 \\ \hline \textbf{1} & 1 & 0 & 0 & 0 & 0 \\ \hline \end{array}}$$

since carry is generated, it has to be neglected & result is taken as shown below

ie. $(10000)_2 = (16)_{10}$

(iii) Subtract 31 from 15

$$\begin{array}{r}
 15 \\
 - 31 \\
 \hline
 - 16
 \end{array}
 \rightarrow \text{binary equivalent is : } 01111 \rightarrow \text{write the positive number as it is}$$

$$\rightarrow \text{binary equivalent is : } 11111 \rightarrow \text{write the 2's complement of the negative number}$$

$$\begin{array}{r}
 \text{ie. } 15 \\
 - 31 \\
 \hline
 - 16
 \end{array}
 \rightarrow \text{binary equivalent is} \rightarrow \boxed{\begin{array}{|c|c|c|c|c|c|} \hline & 0 & 1 & 1 & 1 & 1 \\ \hline + & 0 & 0 & 0 & 0 & 1 \\ \hline \textbf{1} & \textbf{0} & \textbf{0} & \textbf{0} & \textbf{0} & \textbf{0} \\ \hline \end{array}}$$

since carry is not generated, complement the resultant number, add 1 to it & attach a negative sign as shown below

Resultant number

ie. Resultant number is $(10000)_2$ & the answer will be complement of the resultant number plus 1 , with a negative sign,

ie. $(\square 01111 \square 1)_2 = (\square 10000)_2 = (\square 20)_{10}$

(iv) Subtract 45 from 25

$$\begin{array}{r}
 25 \\
 - 45 \\
 \hline
 - 20
 \end{array}
 \rightarrow \text{binary equivalent is : } 011001 \rightarrow \text{write the positive number as it is}$$

$$\rightarrow \text{binary equivalent is : } 101101 \rightarrow \text{write the 1's complement of the negative number}$$

$$\begin{array}{r}
 \text{ie. } 25 \\
 - 45 \\
 \hline
 - 20
 \end{array}
 \rightarrow \text{binary equivalent is} \rightarrow \boxed{\begin{array}{|c|c|c|c|c|c|c|} \hline & 0 & 1 & 1 & 0 & 0 & 1 \\ \hline + & 0 & 1 & 0 & 0 & 1 & 1 \\ \hline \textbf{1} & \textbf{0} & \textbf{1} & \textbf{1} & \textbf{0} & \textbf{0} & \textbf{0} \\ \hline \end{array}}$$

since carry is not generated, complement the resultant number, add 1 to it & attach a negative sign as shown below

Resultant number

ie. Resultant number is $(101100)_2$ & the answer will be complement of the resultant number plus 1 , with a negative sign,

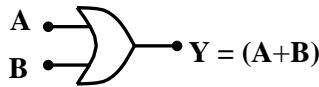
ie. $(\square 010011 \square 1)_2 = (\square 010100)_2 = (\square 20)_{10}$

LOGIC GATES

(1) **OR-GATE** :- Figure shows the logic circuit of a 2 input OR gate. The 2 inputs result in 4 input combinations of 0s & 1s. The operating conditions of the 4 combinations is summarized in the following truth table :-

A	B	$Y = (A+B)$
0	0	0
0	1	1
1	0	1
1	1	1

Logic symbol for OR Gate



The OR operation is represented by the operator “ + ”

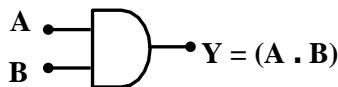
*Input A = Logic 0 or 1
Input B = Logic 0 or 1*

*Logic – 0 = 0 Volt
Logic – 1 = 5 Volts*

(2) **AND-GATE** :- Figure shows the logic circuit of a 2 input AND gate. The 2 inputs result in 4 input combinations of 0s & 1s. The operating conditions of the 4 combinations is summarized in the following truth table :-

A	B	$Y = (A \cdot B)$
0	0	0
0	1	0
1	0	0
1	1	1

Logic symbol for AND Gate :



The AND operation is represented by the operator “ . ”

*Input A = Logic 0 or 1
Input B = Logic 0 or 1*

*Logic – 0 = 0 Volt
Logic – 1 = 5 Volts*

(3) **NOT-GATE** :- Figure shows the logic circuit of a NOT gate (Inverter). It is single input circuit in which the output is a complement of the input ie. if the input is logic-1 the output will be logic-0 & vice versa. As it has a single input, there are only two possible inputs 0 & 1. The NOT gate operation is explained for these two input combinations.

A	$Y = (A)'$
0	1
1	0

Logic symbol for NOT Gate :



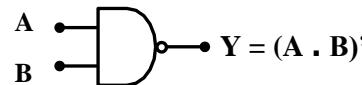
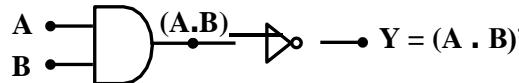
The NOT operation is represented by the operator “ , ”

Input A = Logic 0 or 1

*Logic – 0 = 0 Volt
Logic – 1 = 5 Volts*

(4) **NAND-GATE** :- Figure shows the logic circuit of a 2 input NAND gate. A 2-input NAND gate is realised using an AND gate & a NOT gate. It is actually a combination of a two input AND Gate & a NOT Gate as shown in the logic circuit. It is also called a Negated AND gate (AND gate followed by a NOT gate). The logic symbol for a 2-input NAND gate is also shown along with the truth table .

A	B	$Y = A \cdot B$	$Y = (A \cdot B)'$
0	0	0	1
0	1	0	1
1	0	0	1
1	1	1	0

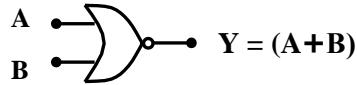
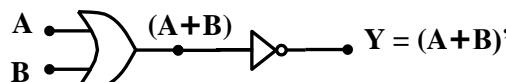


*Input A = Logic 0 or 1
Input B = Logic 0 or 1*

*Logic – 0 = 0 Volt
Logic – 1 = 5 Volts*

(5) **NOR-GATE** :- Figure shows the logic circuit of a 2 input NOR gate. A 2-input NOR gate is realised using an OR gate & a NOT gate. It is actually a combination of a two input OR Gate & a NOT Gate as shown in the logic circuit. It is also called a Negated OR gate (OR gate followed by a NOT gate). The logic symbol for a 2-input NOR gate is also shown along with the truth table .

A	B	$Y = A+B$	$Y = (A+B)'$
0	0	0	1
0	1	1	0
1	0	1	0
1	1	1	0

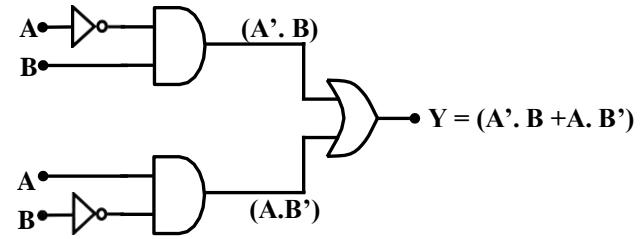
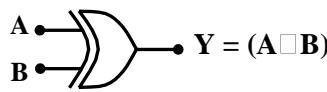


*Input A = Logic 0 or 1
Input B = Logic 0 or 1*

*Logic – 0 = 0 Volt
Logic – 1 = 5 Volts*

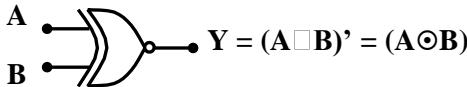
(6) **EXCLUSIVE-OR GATE [EX-OR GATE]** :- The Exclusive-OR gate can be derived using the basic gates ie. AND, NOT & OR gates, or the universal gates ie. NAND or NOR gates. The basic gate realisation for a 2-input EX-OR gate along with the logic symbol & truth table is as shown.

A	B	$Y = (A \square B)$
0	0	0
0	1	1
1	0	1
1	1	0



The EX-OR operation is represented by the operator “ \square ” & output equation is given by : $Y = A \square B = (A' \cdot B + A \cdot B')$

(7) **EXCLUSIVE NOR – GATE [Ex – NOR GATE]** :-



A	B	$Y = (A \square B)$	$Y = (A \square B)' = (A \odot B)$
0	0	0	1
0	1	1	0
1	0	1	0
1	1	0	1

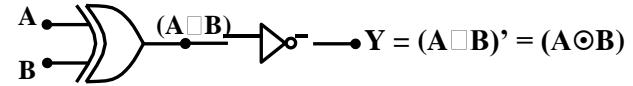


Figure shows the logic symbol of a 2 input EX-NOR gate .It is a combination of a two input EX-OR Gate & a NOT-Gate. It is also called a Negated EX-OR gate (EX-OR gate followed by a NOT gate). The realization of a 2-input EX-NOR gate using an EX-OR gate & a NOT gate along with the truth table is also shown .

The EX-NOR operation is represented by the operator “ \odot ”
The output equation is given by : $Y = A \odot B = (A \cdot B + A' \cdot B')$

DE MORGAN'S THEOREM : Statement of De Morgan's Theorem :

I theorem : The complement of the sum is equal to the product of the complements.
ie. $(A+B)' = (A' \cdot B')$

II theorem : The complement of the product is equal to the sum of the complements.
ie. $(A \cdot B)' = (A' + B')$

Note : Here the **sum** & **product** refer to the **Boolean sum** & **product** ie. **OR** & **AND** respectively

Proof of De Morgan's I Theorem :

A	B	$(A+B)$	$(A+B)'$	A'	B'	$(A' \cdot B')$
0	0	0	1	1	1	1
0	1	1	0	1	0	0
1	0	1	0	0	1	0
1	1	1	0	0	0	0

↑ IDENTICAL ↑

Since entries in the two columns shown
are identical, the theorem is proved

Proof of De Morgan's II Theorem :

A	B	$(A \cdot B)$	$(A \cdot B)'$	A'	B'	$(A' + B')$
0	0	0	1	1	1	1
0	1	0	1	1	0	1
1	0	0	1	0	1	1
1	1	1	0	0	0	0

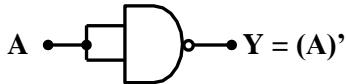
↑ IDENTICAL ↑

Since entries in the two columns shown
are identical, the theorem is proved

UNIVERSAL LOGIC GATES :- A universal logic gate can be used to realize all the basic & derived gates (ie. OR, AND, NOT etc.) . Practically it is observed that NAND & NOR gates function as universal gates ie. it is possible to realize all basic & derived gates using NAND & NOR gates.

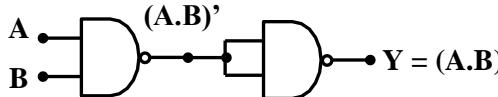
(1) NAND – GATE AS UNIVERSAL GATE :-

(i) Realisation of NOT gate :-



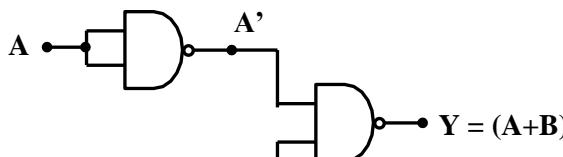
$$Y = (A \cdot A)' = (A)'$$

(ii) Realisation of AND gate :-



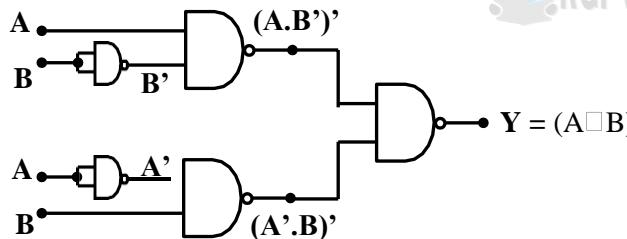
$$Y = [(A \cdot B)']' = (A \cdot B)$$

(iii) Realisation of OR gate :-

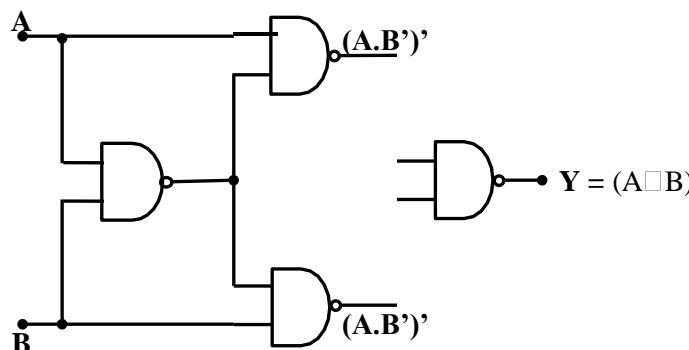


$$Y = (A' \cdot B')' = (A + B)$$

(iv) Realisation of Ex - OR gate :-



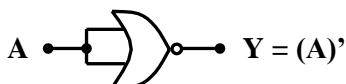
$$Y = (A \cdot B)' + (A' \cdot B)$$



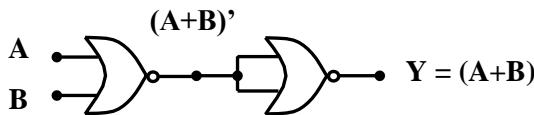
$$Y = (A \cdot B)' + (A' \cdot B)$$

(2) NOR – GATE AS UNIVERSAL GATE :-

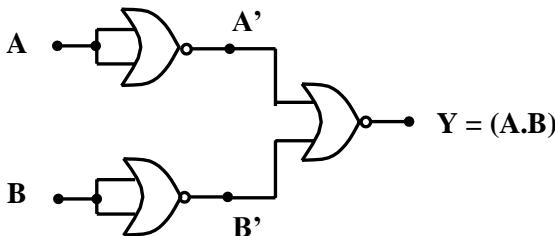
(i) Realisation of NOT gate :-



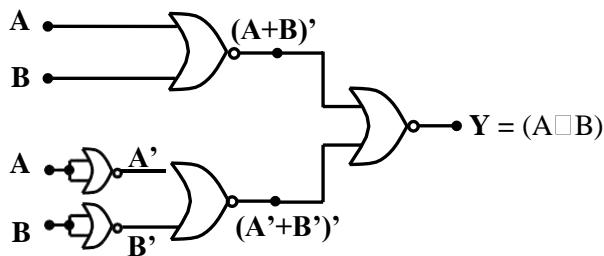
$$Y = (A + A)' = (A)'$$

(ii) Realisation of OR gate :-

$$Y = [(A + B)']' = (A + B)$$

(iii) Realisation of AND gate :-

$$Y = (A' + B')' = (A \cdot B)$$

(iv) Realisation of Ex - OR gate :-

$$Y = (A \cdot B') + (A' \cdot B)$$

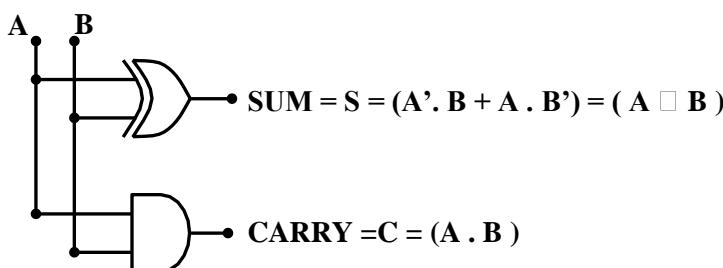
(1) HALF ADDER :- It is a logic circuit used to add 2 one bit binary numbers. A half adder circuit has two inputs & two outputs (sum & carry) . The addition of 2 bits can be shown using the following truth table:

A	B	SUM(S)	CARRY(C)
0	0	0	0
0	1	1	0
1	0	1	0
1	1	0	1

The logic circuit for a half adder is realized using the Boolean expression obtained from the truth table :-

- (i) $\text{Sum} = S = (A' \cdot B + A \cdot B') = (A \oplus B)$
- (ii) $\text{Carry} = C = (A \cdot B)$

The Half Adder circuit is therefore realized as shown below :



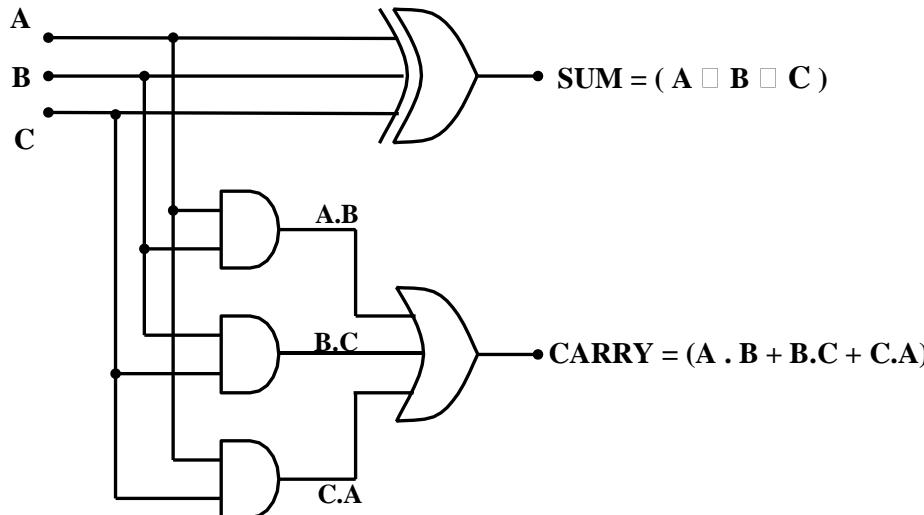
(2) FULL ADDER :- The Half adder circuit can be used to add 2 one bit binary numbers effectively, but when multi bit numbers are to be added then the carry bit that is generated should also be taken care of. This carry bit has to be added to the existing two input bits, which means this circuit would require 3 inputs, ie. two input terminals to add the actual input bits & an additional input terminal to handle the carry bit generated from the previous addition. This is done using a Full adder circuit which is realized using 2 Half adders & a single OR – Gate as shown . The logic circuit for a Full adder is realized using the Boolean expression obtained from the truth table which is shown:-

A	B	C	SUM(S)	CARRY(C)
0	0	0	0	0
0	0	1	1	0
0	1	0	1	0
0	1	1	0	1
1	0	0	1	0
1	0	1	0	1
1	1	0	0	1
1	1	1	1	1

(i) $\text{SUM} = (A' \cdot B' \cdot C + A' \cdot B \cdot C' + A \cdot B' \cdot C' + A \cdot B \cdot C)$
 $= (A \cdot B' \cdot C' + A' \cdot B \cdot C' + A \cdot B' \cdot C + A \cdot B \cdot C)$
 $\text{SUM} = (A \oplus B \oplus C)$

(ii) $\text{CARRY} = (A' \cdot B \cdot C + A \cdot B' \cdot C + A \cdot B \cdot C' + A \cdot B \cdot C)$
 $= BC(A' + A) + A \cdot B' \cdot C + A \cdot B \cdot C'$
 $= B \cdot C + A \cdot B' \cdot C + A \cdot B \cdot C' = B(C + C' \cdot A) + A \cdot B' \cdot C$
 $= B(C + A) + A \cdot B' \cdot C = B \cdot C + B \cdot A + A \cdot B' \cdot C$
 $= C(B + B' \cdot A) + B \cdot A = C \cdot (B + A) + B \cdot A$
 $\text{CARRY} = A \cdot B + B \cdot C + C \cdot A$

The Full adder circuit is therefore realized as shown :



A Full adder can also be realized using two half adders & a single 2 – input OR – gate as shown :

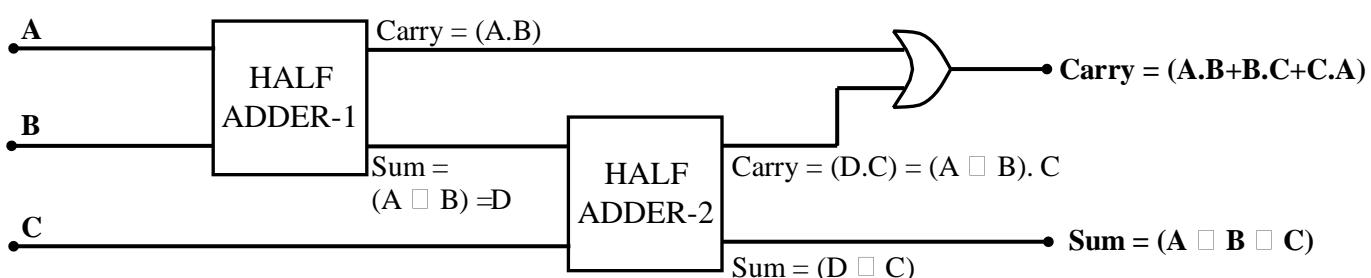
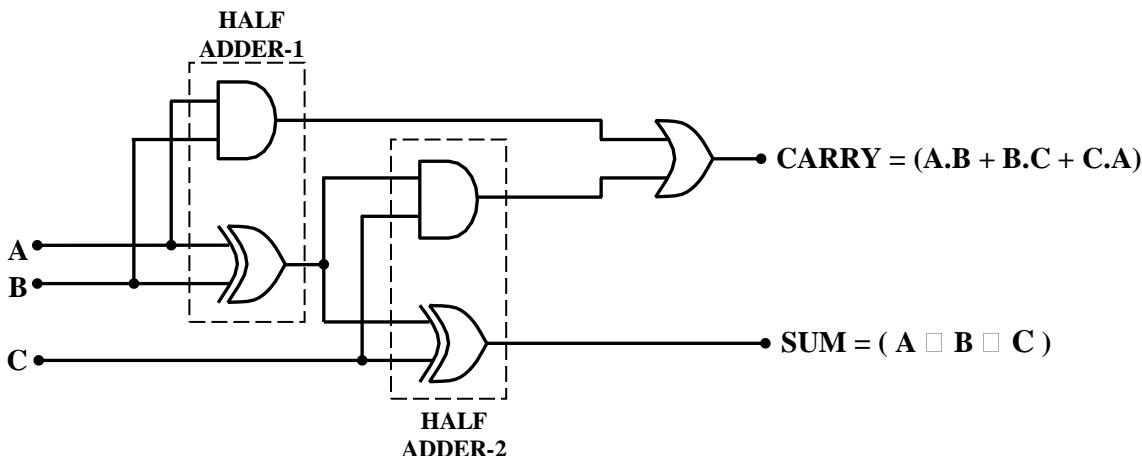
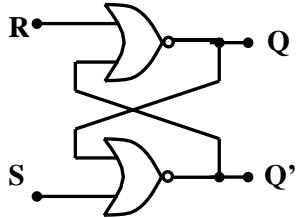


Figure below shows a Full adder realized using two half adders consisting of 2 input Ex-OR gates, 2-input AND gates & a 2-input OR – gate :



R-S FLIP FLOP : A flip flop is a basic memory element (data storage element). A flip flop is realised using a group of logic gates. A NAND gate or a NOR gate individually cannot act as a storage element but when two gates are cross coupled with feed back then they can work as storage or memory elements. Such cross coupled NAND gates or NOR gates with feedback are known as flip flops. A flip flop is a bistable electronic circuit that has two stable states , which means the flip flop output will permanently remain either 0 (low) or 1(high) until it is forced to change its state by an external trigger. A flip flop circuit will have two outputs, one is the Q output & the other is the Q' output which will always be the complement of the Q output , ie Q & Q' are always complementary to each other. Flip flops can be realised using two cross coupled inverters, hence we can use a NOR gate inverter or a NAND gate inverter as shown.

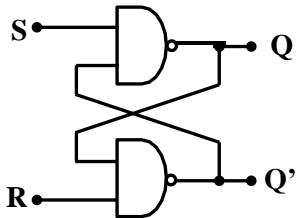


S	R	Q (Output)
0	0	No Change
0	1	0 (Reset)
1	0	1 (Set)
1	1	Race

No Change or Last State or Memory State

Race or Invalid or Not Allowed or ? State

The truth table shown for a NOR gate inverter flip flop is similar to that of a transistor flip flop .

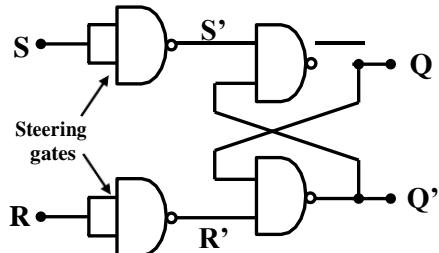


S	R	Q (Output)
0	0	Race
0	1	1 (Set)
1	0	0 (Reset)
1	1	No Change

Race or Invalid or Not Allowed or ? State

No Change or Last State or Memory State

The truth table shown for a NAND gate inverter flip flop is the inverted form of that shown for a NOR gate flip flop, hence inverters or steering gates are used to drive the inputs to the gates as shown :



S	R	Q (Output)
0	0	No Change
0	1	0 (Reset)
1	0	1 (Set)
1	1	Race

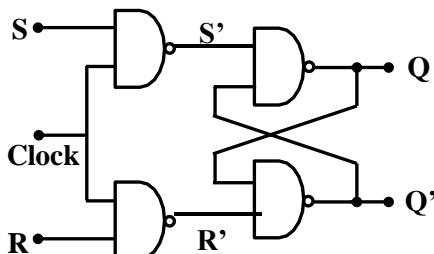
No Change or Last State or Memory State

Race or Invalid or Not Allowed or ? State

The truth table shown for a NAND gate inverter flip flop with steering gates is similar to that shown for a transistor flip flop, hence inverters or driving gates are used to realize the desired practical R-S flip flop.

In order to overcome the RACE problem in R-S flip flops the J-K flip flop is used.

CLOCKED R-S FLIP FLOP :



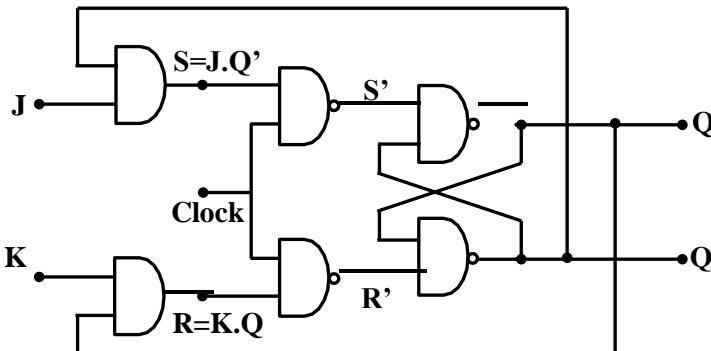
Clock	S	R	Q (Output)
X (0 or 1)	0	0	No Change
1	0	1	0 (Reset)
1	1	0	1 (Set)
1	1	1	Race

X – Don't Care Condition,
ie. Clock is either 0 or 1

The clock signal is also known as the enabling signal which makes the logic circuit perform the required operation.If clock = 0 then the logic circuit will not respond to the input signals ie. the circuit output will remain unaltered. Only when the clock = 1(rising or falling edge) the logic circuit is enabled & will respond to the applied input signals.

J-K FLIP FLOP : A J-K flip flop is realized using a clocked S-R flip flop and two AND gates with appropriate feed back as shown in figure. The problem with the R-S flip flop is that it exhibits the RACE condition when both S & R are high ie when both are logic-1. This condition is a logically unpredictable state. The J-K flip flop eliminates the unpredictable condition that occurs in the S-R flip flop and hence can be practically used in logic circuits. The J input is analogous to the S input & the K input is analogous to the R input. This means that when J=1 & K=0 , the J-K flip

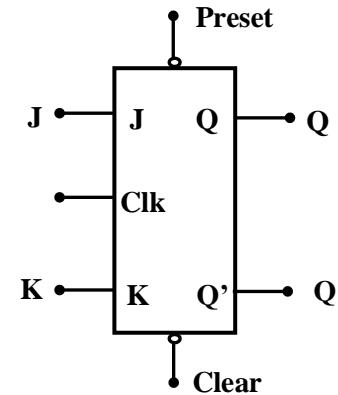
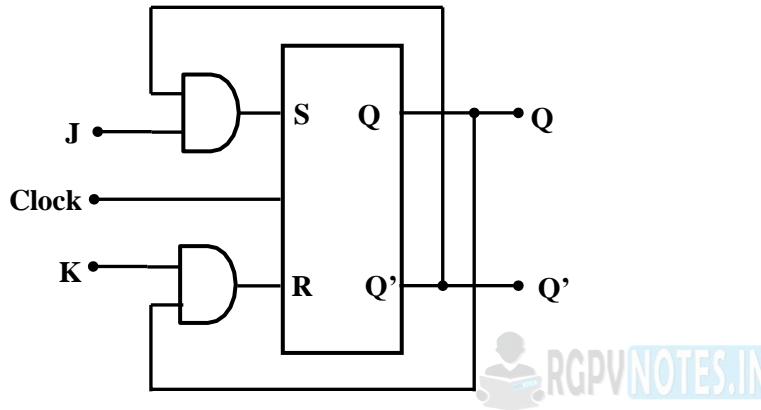
will Set , ie. $Q=1$, and when $J=0 \& K=1$, the J-K flip flop will Reset, ie. $Q=0$. As usual there will be no change in the output condition when $J=K=0$. However the most important change when compared to the S-R flip flop is that the J-K flip flop will complement its output condition when $J=K=1$ with the clock high. The operation of a J-K flip can be clearly understood from the truth table given below.



Realisation using S-R flip flop & AND gates

Input $S = J \cdot Q'$
Input $R = K \cdot Q$

Logic Symbol of J-K flip flop



Clk	J	K	Q_n	$Q_{n'}$	$S = J \cdot Q_n$	$R = K \cdot Q_n$	Output (Q_{n+1})	Remarks
1	0	0	0	1	0	0	0	= Q_n ie. No Change or Last State or Memory
1	0	0	1	0	0	0	1	
1	0	1	0	1	0	0	0	= 0 , ie. Reset (Make $Q = 0$)
1	0	1	1	0	0	1	0	
1	1	0	0	1	1	0	1	= 1 , ie. Set (Make $Q = 1$)
1	1	0	1	0	0	0	1	
1	1	1	0	1	1	0	1	= Q_n' ie. Toggle or Complement or Switch to opposite state
1	1	1	1	0	0	1	0	

Q_n represents the Present State ; Q_{n+1} represents the Next State ie. the state of the output after the clock pulse is applied.

Race Around Condition in J-K Flip Flop : By using two AND gates & appropriate feed back the RACE problem that existed in the S-R flip flop could be eliminated in a J-K flip flop. However there is a problem of an unpredictable state occurring in the J-K flip flop also . Due to this problem the Q output will start oscillating between the 0 (low) & 1 (high) states .The output condition therefore could be either 0 or 1. This problem is known as the Race around Condition. The race around condition in a J-K flip flop occurs when $J=1$, $K=1$ and the clock is also =1, with the clock pulse width “ t_p ” greater than the propagation delay “ Δt ” of the gates. We assume that the inputs of the J-K flip flop do not change during a clock pulse , but due to the feed back they change when the clock remains high (1), hence the output condition starts oscillating between the low & high states. This problem can be avoided by making the clock pulse width less than the propagation delay of the gates, but practically this is difficult because the propagation delay is very small, hence the problem of Race Around Condition is overcome using a Master-Slave J-K flip flop. In this flip flop the input conditions do not change when the clock remains high, hence the output does not oscillate.

PROPERTIES OF SEMICONDUCTOR MATERIAL :- A semiconductor is a material which exhibits the following properties :-

- 1) It has a resistivity lying between that of a conductor and an insulator.
- 2) It is tetravalent.
- 3) It exhibits negative temperature co-efficient of resistance .
- 4) It exhibits crystalline structure.
- 5) Its conductivity increases when doped with trivalent or pentavalent atoms.

Silicon (atomic no. 14) and Germanium (atomic no. 32) are the two most important semiconductor materials.

INTRINSIC SEMICONDUCTOR :- A semiconductor material in its purest form is known as an intrinsic semiconductor. An intrinsic semiconductor behaves as an insulator at 0 K but acts as a conductor at 300 K (room temperature). At room temperatures due to the thermal generation of electron-hole pairs , free electrons & holes are generated in equal numbers , these mobile charges help in the conduction of current in an intrinsic semiconductor. Since electron-hole pairs that are responsible for conduction of current in an intrinsic semiconductor are internal to the semiconductor crystal , the material is known as an intrinsic semiconductor .

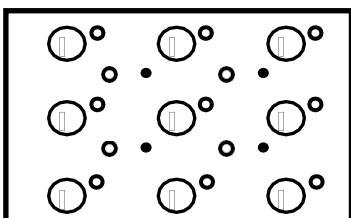
FREE ELECTRONS OR CONDUCTION ELECTRONS :- When external energy is supplied to a semiconductor crystal in the form of light or heat (increase in temperature), some covalent bonds break and produce free electrons. Every free electron has an associated vacant site (hole) in the covalent bond .These free electrons are not under the control of any of the nuclei within the crystal.. Since free electrons take part in the conduction of current , they are also known as conduction electrons. Conduction electrons or free electrons have energy levels much higher than valence electrons and take part in the conduction of current in a semiconductor.

VALENCE ELECTRONS OR BOUND ELECTRONS :- The outer most orbit electrons or valence electrons are shared by the neighboring semiconductor atoms to form covalent bonds in a crystal. A valence electron is always associated with a particular nuclei and is under it's control, hence a valence electron is also known as a bound electron. A valence electron by itself cannot take part in the conduction of current. A valence electron will take part in the conduction of current only when there is hole movement, in other words hole movement is actually the movement of valence electrons in the valence band. At 0 K all the electrons in a Silicon crystal exist as valence electrons (ie. there are no free electrons or holes) hence there is no current conduction & Silicon behaves as an insulator.

DOPING :- The process of adding a calculated quantity ($1:10^8$) of trivalent or pentavalent atoms to an intrinsic semiconductor is known as doping. Doping helps in generating a single type of charge carrier (either free electrons or holes). Doping thus increases the conductivity of a semiconductor at room temperatures.

EXTRINSIC SEMICONDUCTOR :- An extrinsic semiconductor is obtained by doping an intrinsic semiconductor with trivalent or pentavalent impurity atoms . Depending upon the valency of the impurity atoms added we obtain either p-type or n-type extrinsic semiconductor.

P-TYPE EXTRINSIC SEMICONDUCTOR :-

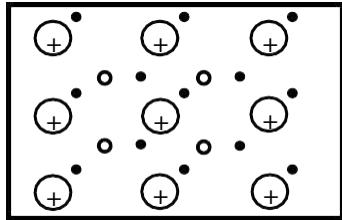


- Hole (Majority Carrier)
- Free electron (Minority Carrier)
- Negative Immobile ion

A p-type extrinsic semiconductor is obtained by adding trivalent impurity atoms (Boron, Aluminum, Gallium or Indium) to an intrinsic semiconductor in the ratio of $1:10^8$.

The three valence electrons of aluminum atom form three co-valent bonds with the three neighboring semiconductor atoms . The absence of a fourth valence electron creates a vacant site (hole) in one of the covalent bonds . The hole has a positive charge & is capable of accepting an electron from the semiconductor crystal to become a negative immobile ion. Thus a trivalent impurity atom is known as an acceptor impurity . Every trivalent atom added to a semiconductor crystal gives rise to a hole and a negative immobile ion. The conductivity of a p-type semiconductor is directly proportional to the doping density. In a p-type semiconductor, holes are in excess when compared to conduction electrons , hence holes are known as majority carriers and conduction or free electrons are known as minority carriers . The total current flowing through a p-type semiconductor is the sum of free electron and hole currents. Since majority of this current is due to holes (positive charges) the material is known as a p-type extrinsic semiconductor . A p-type semiconductor is electrically neutral because total number of negative charges is exactly equal to the total number of positive charges.

N -TYPE EXTRINSIC SEMICONDUCTOR :- A n-type semiconductor is obtained by adding pentavalent impurity atoms (phosphorus , arsenic or antimony) to an intrinsic semiconductor in the ratio of $1:10^8$.The four valence electrons of phosphorus form four co-valent bonds with the four neighboring semiconductor atoms. The fifth valence electron cannot form a co-valent bond hence remains free. This free electron helps in the conduction of current. The pentavalent impurity atom loses one electron and becomes a positive immobile ion . When a pentavalent atom is added to an intrinsic semiconductor one electron is donated into the crystal, hence a pentavalent



- Free electron (Majority Carrier)
- Hole (Minority Carrier)
- Positive Immobile ion

impurity atom is also known as a donor impurity. Every pentavalent impurity atom added gives rise to a free electron and a positive immobile ion. The conductivity of a n-type semiconductor is directly proportional to the doping density. In a n-type semiconductor, electrons are in excess when compared to holes, hence electrons are majority carriers and holes are minority carriers. Since majority of the current is carried by negative charges (electrons), the material is known as n-type semiconductor. A n-type semiconductor is electrically neutral because the total number of positive charges is exactly equal to the total number of negative charges in the crystal

P-N JUNCTION DIODE :- Figure shows the construction and circuit symbol of a p-n junction diode. The arrow mark in the diode symbol indicates the direction of flow of conventional current. A p-n junction diode is fabricated using a single semiconductor crystal, in which one half is doped with p-type impurity, while the other half is doped with n-type impurity. A p-n junction diode is basically a unidirectional device (ie. it allows current to flow in one direction & blocks it in the other direction). A p-n junction diode is a high speed electronic switch & is widely used in rectifier circuits.

The operation of a semiconductor diode is studied under the following three conditions :-

(1) Unbiased condition :-

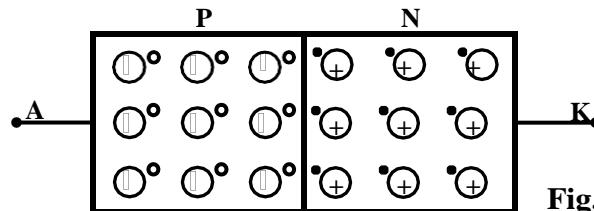


Fig. 1

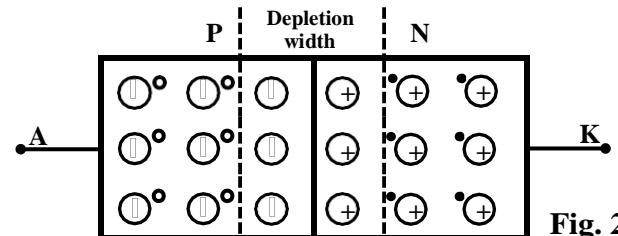


Fig. 2

Fig.2 shows a p-n junction diode under unbiased condition, ie. both the anode & cathode are at the same potential or both are at zero potential. For simplicity only impurity atoms are shown (semiconductor atoms are not shown) because for every impurity atom, there will be 10^8 semiconductor atoms (because doping density is 1:10⁸).

The p-region has negative immobile ions and their corresponding holes as the majority carriers, while the n-region has positive immobile ions and their corresponding free electrons as the majority carriers. Thermally generated electron-hole pairs are also not shown for simplicity. At the instant of junction formation, the p-material has excess holes and the n-material has excess electrons as shown in Fig.1. and the depletion region does not exist. As soon as the p & n regions are formed, electrons on the n-side recombine with holes by crossing onto the p-side of the junction due to diffusion. Soon after recombination both the electrons & the holes disappear and leave behind immobile positive ions on the n-side and immobile negative ions on p-side of the junction as shown in Fig.2. This electric field created by the immobile positive & negative ions on either side of the junction prevents further diffusion of charges. Thus a depletion region (width $\approx 50 \mu\text{m}$) is formed at the junction even under unbiased conditions as shown in Fig.2.

(2) Forward biased condition :-

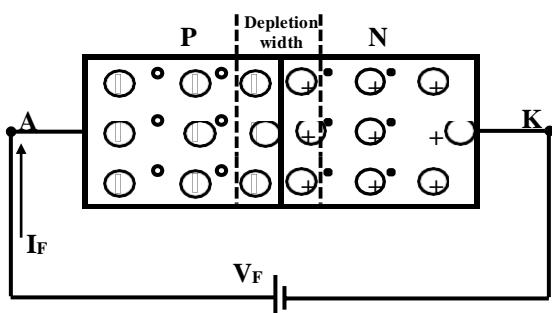


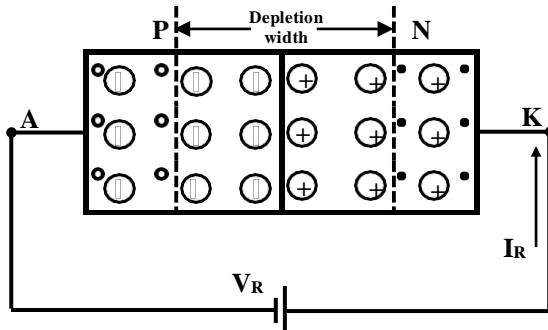
Figure shows a p-n junction diode under forward biased condition (ie. anode is at a higher potential than the cathode). The battery polarity is such that majority carriers in both p & n regions are pushed towards the junction. Since electrons & holes enter the depletion region, it causes a reduction in the depletion width & hence height of the potential barrier. The reduced potential barrier allows a few high-energy electrons on the n-side to cross the junction on to the p-side and constitute a small forward current. As the magnitude of forward bias voltage is increased the depletion width further reduces & thereby further increases the forward current.

The depletion width & the potential barrier reduce to almost zero when the p-n junction is forward biased by a voltage greater than the cut-in voltage V_T (0.7 V for Silicon diode & 0.3 V for Germanium diode).

At voltages greater than V_T , a p-n junction diode acts like a closed switch (offers zero resistance) and a heavy current starts flowing. Practically a very small value resistance is offered due to the existence of the bulk resistance of the semiconductor crystal.

The thermally generated electron-hole pairs present in both p and n regions & the minority carriers also move in the same direction as majority carriers, ie. they also add to the forward current.

(3) Reverse biased condition :-



When a p-n junction diode is reverse biased (ie. anode is at a lower potential than the cathode) a very small reverse current flows through the junction due to a small number of temperature dependent minority charge carriers (electrons in p-region & holes in n-region). This minority current or leakage current is also known as the reverse saturation current & is temperature dependent. The leakage current which has a very small value (1 or 2 μ A) doubles itself for every $10^\circ C$ rise in temperature. The diode therefore offers very high resistance (1 to $2 M\Omega$). This means that the diode acts as an open switch under reverse biased conditions. The battery connection is such that majority carriers in both p and n regions are pulled away from the junction. Thus

both the depletion width and the potential barrier increase under reverse bias conditions. This reverse current flows until the reverse voltage is equal to the junction breakdown voltage. Beyond breakdown voltage, there is a drastic increase in the reverse current which is explained using the avalanche breakdown phenomenon. At voltages beyond V_{BD} , minority carriers (electrons) on the p-side gain sufficiently high velocities to knock out valence electrons from the semiconductor atoms. This is a cumulative effect and is known as ionisation due to collision. A large number of charges are thus available to constitute a large reverse current. If left uncontrolled, this reverse current can cause physical breakdown of the junction. A p-n junction diode under reverse biased condition is therefore operated well within its breakdown voltage.

V-I CHARACTERISTICS OF P-N JUNCTION DIODE :-

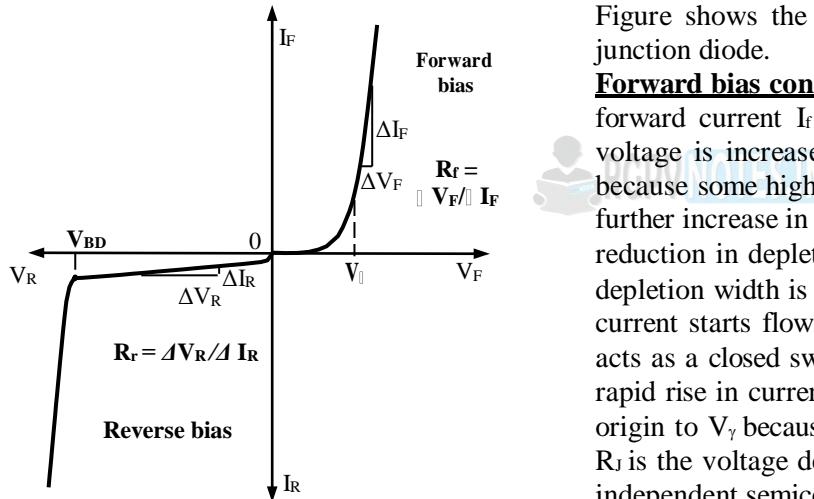


Figure shows the forward & reverse bias characteristics of a p-n junction diode.

Forward bias condition: When the forward bias voltage $V_f = 0$, the forward current I_f is also equal to 0. When the forward bias voltage is increased, current through the diode gradually increases because some high-energy electrons start crossing the junction. Any further increase in V_f causes an increase in forward current due to reduction in depletion width & potential barrier. When $V_f = V_\gamma$, the depletion width is zero & potential barrier is also zero. Now a large current starts flowing through the diode. Thus beyond V_γ the diode acts as a closed switch and offers very low resistance resulting in a rapid rise in current. The characteristics will be non-linear from the origin to V_γ because the total diode resistance $R_T = (R_J + R_B)$, where R_J is the voltage dependent junction resistance and R_B is the voltage independent semiconductor crystal bulk resistance.

The non-linearity in the characteristics from origin to V_γ is because of the junction resistance, which is reducing with an increase in voltage. The characteristic is linear beyond V_γ because the junction resistance becomes zero after V_γ & it is only the voltage independent bulk resistance R_B that remains. Thus the diode starts behaving as a closed switch only beyond V_γ with a very low value of forward resistance R_f .

Reverse bias condition :- When a p-n junction is reverse biased, a very small leakage current flows due to a very small number of temperature dependent minority carriers. The leakage current I_R is also known as reverse saturation current or minority current. This small current continues to flow until the applied reverse voltage is equal to the breakdown voltage V_{BD} . Beyond V_{BD} there is rapid increase in the leakage current due to Avalanche breakdown phenomenon. At voltages beyond V_{BD} , minority electrons on the p-side of the junction gain sufficiently high velocities to knock out valence electrons from the semiconductor atoms within the crystal. This is a cumulative process & a large number of charges are made available to cause a large value of leakage current as shown in the characteristics. This phenomenon is also known as ionization due to collision. Hence a p-n junction diode under reverse bias condition, is operated well within its breakdown voltage if it has to work as an open switch.

DEPLETION REGION :- When a p-n junction is formed there is movement of charges across the junction due to diffusion even under unbiased conditions. This results in uncovering of the Donor ions (positive immobile ions) on the n-side & the Acceptor ions (negative immobile ions) on the p-side (Refer to Fig.2 on page-2). This region on either sides of the junction consisting of the uncovered immobile positive & negative ions is known as the Depletion Region. Since this region is depleted of mobile charges ie. there are no mobile charges it is known as the

Depletion region or Transition Region. It is also known as the space charge region because it consists of immobile positive & negative ions. Since this region does not contain mobile charges it behaves as an insulator. The region across the junction which does not have mobile charge carriers is known as the depletion width.

When a p-n junction is forward biased, due to the applied voltage the Depletion Width gets reduced and the diode starts conducting because of the lower value of junction resistance.

When the p-n junction is reverse biased, due to the applied reverse bias voltage the Depletion Width increases and the diode offers a very high resistance to the flow of current due to the increased width of the insulator.

BARRIER POTENTIAL :- The electric field that exists across a p-n junction between the positive immobile ions on the n-side & negative immobile ions on the p-side of the depletion region is known as the Barrier. Uncovered donor & acceptor ions exist on either sides of a p-n junction. These are isolated positive & negative electrical charges which can result in an electric field at the junction. This electric field prevents further diffusion of holes & electrons across the junction under unbiased conditions, ie. it acts as an obstruction or barrier to the movement of electric charges, hence it is known as the potential barrier. The physical distance from one side of the barrier to the other side is known as the barrier width. The difference in potential between the two sides is known as the height of potential barrier. The potential barrier is approximately 0.7V for a Silicon diode & 0.3V for a Germanium diode.

DIODE APPLICATIONS :

Diode clipping circuits :

A clipper circuit is a diode network which can clip off a portion of the input signal without distorting the remaining waveform. Clipper circuits are also known as Slicers or Limiters. Depending on the diode connection in the circuit the positive or negative side of the input waveform is clipped. If a dc voltage source is connected in series with the diode it is known as a biased clipper if not it is known as an unbiased clipper. If the diode is in parallel with the output terminals the circuit is known as a parallel clipper (shunt clipper) and if the diode is in series with the output terminals it is known as a Series clipper. The reference voltage at which the waveform has to be clipped should always be less than the maximum or peak voltage of the input signal (ie. $V_R < V_m$).

In all clipper circuits it is assumed that the diode used for clipping is an ideal diode ie. its cut-in voltage is zero (ie. $V_g = 0$) & the forward resistance is also zero(ie. $R_f = 0$).

Negative wave clipper:

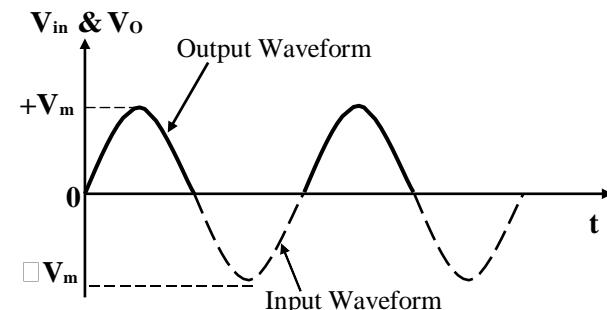
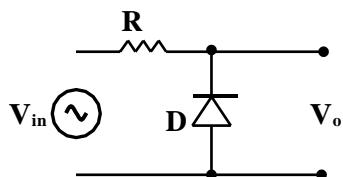


Figure shows the circuit diagram of a negative wave clipper. The diode D is assumed to be an ideal clipping device ie. its cut-in voltage $V_g = 0$ & the forward resistance $R_f = 0$.

During the positive half cycle diode D gets reverse biased and acts like an open switch (infinite resistance) . Hence the output wave form follows the input waveform.

During the negative half cycle diode D gets forward biased and acts like a closed switch (zero resistance) . Hence the output voltage across a zero resistance will be zero.

In this circuit only the negative half cycles are clipped while the positive half cycles appear at the output.

Diode Clamping Circuits :

The circuit is also known as DC level shifter or DC level restorer. A diode clamping circuit is used to fix an AC signal to any desired DC level (zero, positive or negative) without changing the shape or amplitude of the input signal. Both the positive & negative peaks can be clamped to any desired DC level. It is important that the (R.C) time constant of the circuit is very large when compared to the time period of the input waveform. This condition will keep the output waveform undistorted (ie. the square wave will have constant amplitude during the entire half

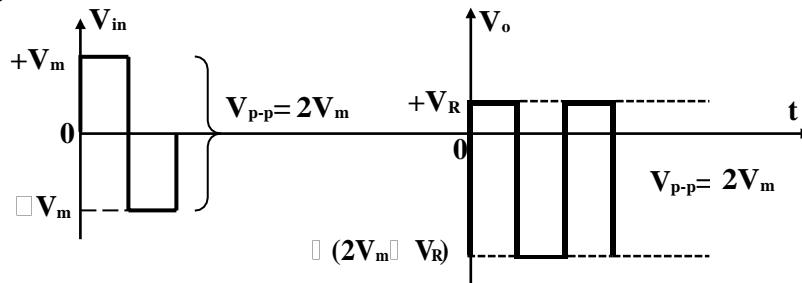
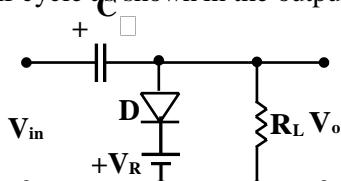
cycle). The reference voltage to which the waveform has to be fixed or clamped should always be less than the maximum or peak voltage (V_m) of the input signal.

In all clamper circuits it is assumed that the diode used for clamping is an ideal diode ie. its cut-in voltage is zero (ie. $V_g = 0$) & the forward resistance is also zero (ie. $R_f = 0$).

Positive peaks clamped to positive reference voltage :

Figure shows the circuit diagram which clamps the positive peaks of the input waveform to a positive reference voltage level. The diode D is assumed to be an ideal clamping device ie. its cut-in voltage $V_g = 0$ & the forward resistance $R_f = 0$.

During the positive half cycle , diode D gets forward biased and capacitor C charges to a voltage equal to $(V_m - V_R)$ with a polarity as shown in the circuit. Since the diode is a closed switch (zero cut-in voltage & zero forward resistance), the output voltage will be equal to the reference voltage($+V_R$ volts) during the entire positive half cycle as shown in the output waveform.



During the negative half cycle, the input voltage will be in series with the capacitor voltage which is already charged to $(V_m - V_R)$, the diode is kept reverse biased by the sum of the input & capacitor voltages. The diode now behaves as an open switch and the output voltage will be equal to the sum of the input & capacitor voltages ie. equal to $[-(2.V_m - V_R)]$. Thus in the output voltage waveform the positive peaks appear to be clamped to a positive reference voltage level V_R during the entire negative half cycle. After clamping it is observed that the output peak to peak voltage swing remains unchanged at a value equal to $(2.V_m)$.

Half wave rectifier :

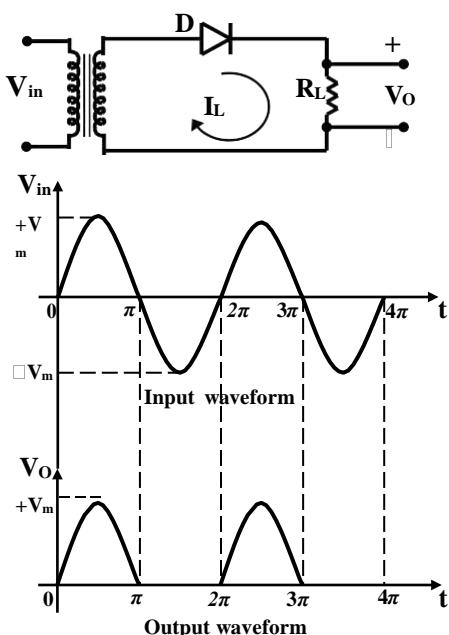


Figure shows the circuit diagram of a half wave rectifier. It consists of a silicon diode D connected to the secondary of a step down transformer. The load resistance R_L is connected in series with the diode.

During the positive half cycle (0 to π), the diode D is forward biased and acts like a closed switch offering very low resistance. The load current I_L varies in accordance with the input voltage V_{in} ($I_L \propto V_{in}$) as shown in figure. The output voltage V_o is equal to $(I_L.R_L)$, since R_L is a constant the output voltage $V_o \propto I_L$. Thus during positive half cycle, the output voltage follows the load current & in turn the input voltage V_{in} .

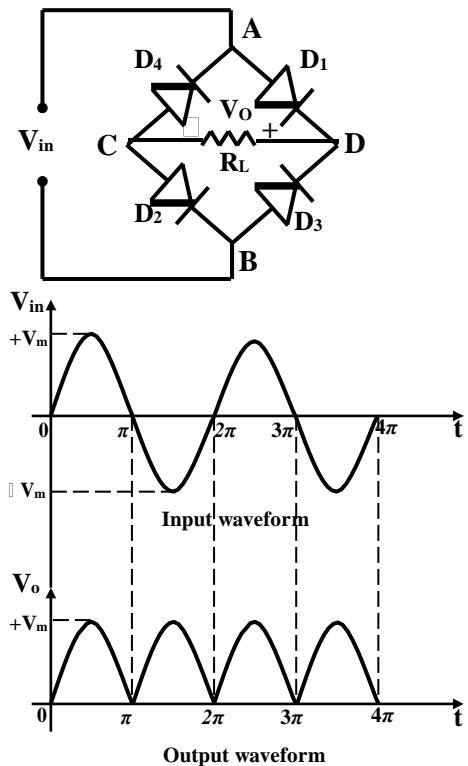
During the negative half cycle(π to 2π), the diode D is reverse biased and acts like an open switch ie. the diode offers very high resistance. Thus the load current I_L is almost zero, so that V_o is zero during the negative half cycle

This cycle repeats and the circuit produces an output only during one half of the input cycle. This circuit is therefore known as a half wave rectifier. The output of this circuit is a pulsating DC whose frequency is equal to the input frequency. For a half wave rectifier we have the following important expressions:

$$I_{DC} = (I_m / \pi) ; \quad I_{RMS} = (I_m/2)$$

$$\% \text{ } \Delta = 40.6 ; \quad \text{Ripple Factor} = 1.21$$

Full wave bridge rectifier:



A full wave bridge rectifier uses four diodes as shown in the circuit diagram. The load resistance R_L is connected between C and D, while the input voltage is applied between points A and B.

During the positive half cycle (0 to π), point A is at a higher potential than B, hence diodes D₁ and D₂ are forward biased and act as closed switches. The current I₁ flows from points A to B through D₁, R_L and D₂. Both D₃ and D₄ are reverse biased during the positive half cycle and act as open switches. Thus the output current during positive half cycle due to diodes D₁ and D₂ flows through R_L and produces an output as shown in the waveforms.

During the negative half cycle (π to 2π), point B is at a higher potential than A, hence D_3 and D_4 are forward biased and act as closed switches. Current I_2 flows from point B to point A through D_3 , R_L and D_4 . Both D_1 and D_2 are reverse biased during the negative half cycle and act as open switches. The output current due to D_3 and D_4 flows through R_L producing an output as shown in the waveforms.

The direction of currents I_1 and I_2 through the load resistance R_L during both the half cycles is identical, hence the output voltage has the same polarity and is plotted in the same quadrant. As the circuit produces an output during both the half cycles, it is known as a full wave rectifier. In this case the output voltage is $+V_m$ if the input voltage has a peak value of V_m . The output frequency is twice the input frequency.

For a Full wave rectifier we have the following important expressions:

$$I_{DC} = (2I_M / \square) \quad ; \quad I_{RMS} = (I_M / \square 2)$$

% = 81.2 ; Ripple Factor = 0.48

Rectifier efficiency :- It is defined as the ratio of DC output power to the AC input power. It is a measure of the AC to DC conversion capacity of the rectifier circuit.

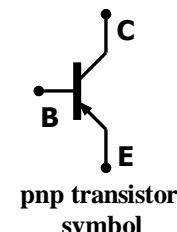
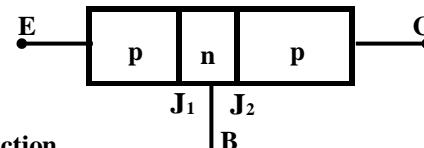
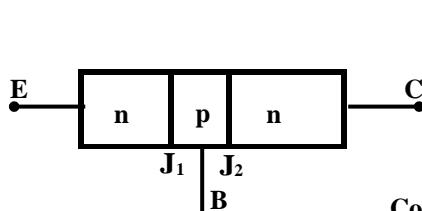
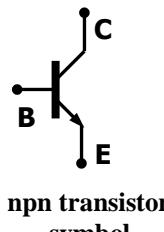
$$\eta = \frac{P_{DC}}{P_{AC}} = \frac{\text{output DC power}}{\text{input AC power}} = \frac{(V_{DC} \cdot I_{DC})}{\{(V_{rms})^2 / (R_f + R_L)\}}$$

Ripple factor :- It is defined as the ratio of RMS value of AC component to the average value of the dc component. It is the measure of the pulsating component in the output.

Ripple factor = ripple voltage / DC value of output

$$\text{Ripple factor} = [\text{RMS value of AC component / DC value of output}] = [(I_{\text{rms}} / I_{\text{dc}})^2 - 1]^{1/2}$$

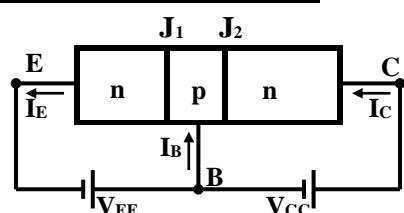
BIPOLAR JUNCTION TRANSISTORS :



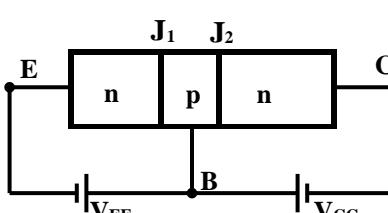
A transistor is fabricated using a single crystal of Germanium or Silicon. It is a 3 terminal device having alternate p and n layers with two junctions J_1 & J_2 . This type of construction results in npn and pnp transistors. In a npn transistor, the p-layer is sandwiched between two n-layers. The first n-layer is the emitter which emits electrons. The n-type emitter layer is heavily doped to provide better injection efficiency. The other n-layer is the collector which collects electrons. The collector region is moderately doped and has a large width to help better heat dissipation. The p-type base layer forms one junction (J_1) with the emitter and another junction (J_2) with the collector layer. The base region is lightly doped and has a narrow width, this helps in reducing recombination in the base and in the process reduces the value of base current & increases the value of collector current.

In a pnp transistor the n-layer is sandwiched between two p-layers. The construction and symbol for npn and pnp transistors is as shown in figure. The arrow mark on the emitter lead indicates the direction of flow of conventional current.

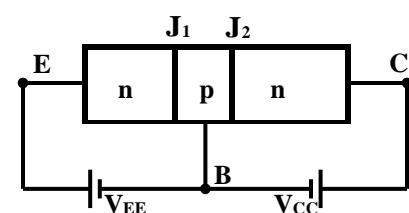
TRANSISTOR BIASING:-



Transistor biased as Amplifier



Transistor biased as closed switch

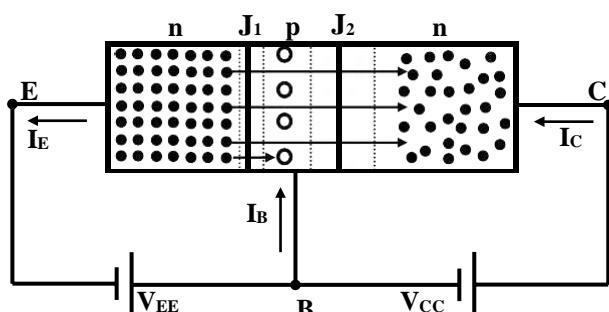


Transistor biased as open switch

If a transistor has to work as an amplifier, the base-emitter junction J_1 must be forward biased and the collector-base junction J_2 must be reverse biased. Transistor biasing is a process of creating an appropriate potential difference across the base-emitter and the collector-base junctions. The base-emitter junction should always be forward biased by a voltage greater than its cut-in voltage (V_{γ}), while the collector-base junction should be sufficiently reverse biased for efficient collection of charges. If these two conditions are satisfied, the transistor provides faithful amplification while operating in the active region (linear region).

A transistor can also operate as a switch (in the non-linear regions). If both the emitter-base & collector-base junctions are forward biased then the transistor will behave as a closed switch offering almost zero resistance(saturation region) . If both the emitter-base & collector-base junctions are reverse biased then the transistor behaves as an open switch offering very high resistance(cut-off region)

TRANSISTOR OPERATION :-



are lost due to recombination in the p-type base region and constitute a small base current. The base current magnitude is kept minimum by using a lightly doped narrow base region. The reverse biased collector-base junction will assist the diffusion of minority carriers (electrons) from base to the collector region. These electrons are then collected by the positive terminal of the battery V_{CC} . Electrons flowing out of the collector constitute a large collector current. Using Kirchoff's law, the fundamental transistor equation can be shown to be : $I_E = I_B + I_C$

Figure shows the battery connections and directions of current in a npn transistor. V_{EE} is the emitter battery, which forward biases the base-emitter junction while V_{CC} is the collector battery which reverse biases the collector-base junction. I_B is the base current, I_C is the collector current and I_E is the emitter current. The forward biased base-emitter junction makes the emitter inject a large number of electrons into the base region. Electrons are minority carriers in the p-type base region, hence they easily diffuse into the collector region. Some electrons

I_B is very small when compared to I_C (3 to 4 % of I_C). Therefore $I_E \approx I_C$, ie. input current = output current. A small reverse current flows through the collector-base junction when the emitter lead is open (when the input current is zero). This reverse current or leakage current is I_{CBO} (collector to base current with emitter open), ie. the output collector current $I_C = I_{CBO}$ when the input current is zero. I_{CBO} is temperature dependent and independent of the applied reverse voltage. In a transistor, a large emitter current flowing through a low resistance input circuit is transferred into a high resistance collector circuit (output circuit), hence it is called a transfer-resistor or a transistor.

TRANSISTOR CONFIGURATIONS :-

A transistor has only 3 leads hence any one of the 3 leads has to be common to the input & output circuits if the transistor is to be considered as a 2-port linear network. Depending on the lead that is common to both the input & output circuits there are three transistor configurations :

- (1) **Common-base configuration or Grounded-base configuration.**
- (2) **Common-emitter configuration or Grounded-emitter configuration.**
- (3) **Common-collector configuration or Grounded-collector configuration.**

The behaviour of a transistor varies greatly with each configuration & can be understood by studying the input & output characteristics in all the 3 configurations.

COMMON BASE CONFIGURATION :-

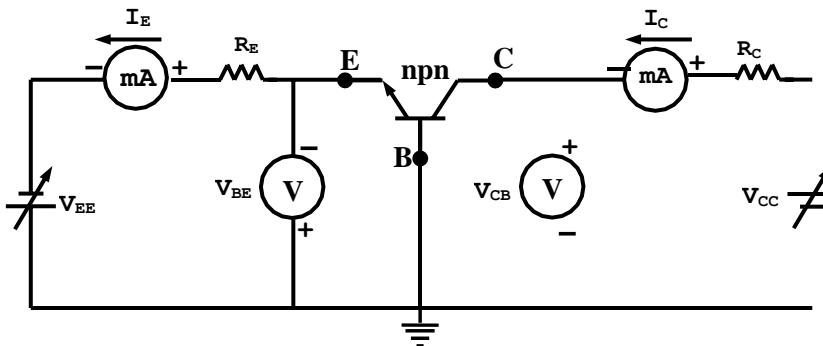


Figure shows the circuit arrangement for obtaining the input and output characteristics of a npn transistor in common-base configuration. V_{EE} is the emitter battery on the input side and R_E is the emitter current limiting resistor. The milliammeter is used to measure the emitter current(input current) while the voltmeter is used to measure the input voltage V_{BE} . V_{CC} is the collector battery on the output side and R_C is the collector resistance.

The milliammeter measures the collector current (output current) while the voltmeter measures the collector-base voltage, V_{CB} (output voltage). Here the base lead is common to both the input and output circuits, hence it is known as the common-base configuration.

Input Characteristic:-

The input characteristics is a plot of Input voltage v/s Input current keeping output voltage constant .

ie. V_{BE} v/s I_E keeping V_{CB} constant.

The input characteristics is obtained by varying V_{BE} in steps and noting down the corresponding values of I_E keeping V_{CB} constant. A family of curves can be obtained for different values of V_{CB} .

The dynamic input resistance r_i is obtained using the relation:

$$r_i = \frac{\partial V_{BE}}{\partial I_E} | V_{CB} = \text{constant.}$$

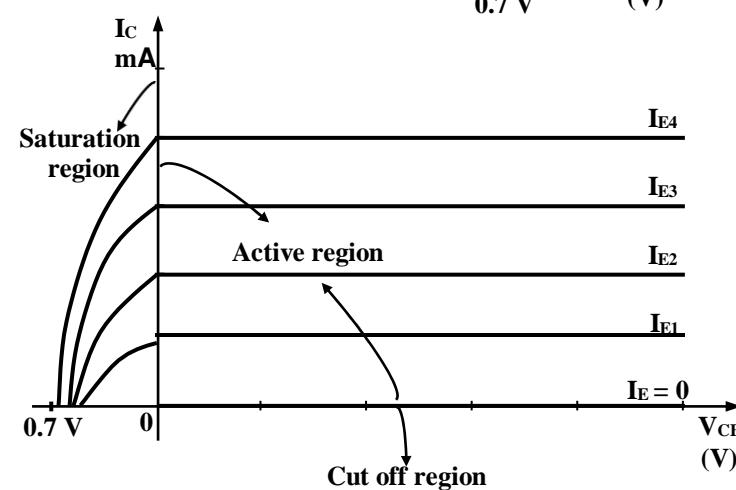
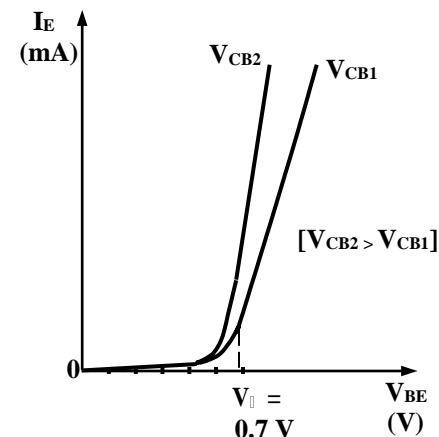
Output Characteristic:-

The output characteristic is a plot of output voltage v/s output current keeping input current constant,

ie. V_{CB} v/s I_C keeping I_E constant.

It is obtained by varying V_{CB} in steps and noting down the corresponding values of I_C keeping I_E constant.

A family of curves are obtained for different values of I_E . $I_C \approx I_E$ because I_B is very small. The slope of the output characteristic is almost zero, this means that the output resistance is very large.



The dynamic output resistance is given by the relation:

$$r_o = \frac{\partial V_{CB}}{\partial I_C} \mid I_E = \text{constant}$$

\square is the current gain of a transistor in common base configuration and is determined using the relation:

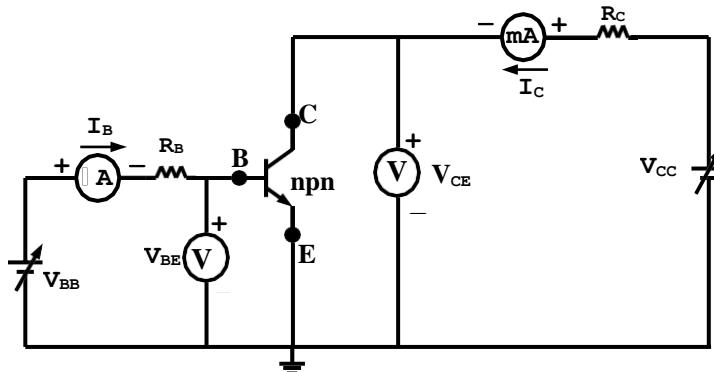
$$\square = \frac{\partial I_C}{\partial I_B} \mid V_{CB} = \text{constant}$$

The input resistance r_i has a very low value (5 to 15 Ω) while the output resistance has a very high value ($\approx 1M\ \Omega$).

The current gain α has a value less than 1 (0.95 to 0.995). The voltage gain is high (150-200).

Applications:- A transistor in common-base configuration is used as a Wide-band amplifier, a constant current source and a buffer amplifier (for impedance matching).

COMMON Emitter CONFIGURATION :



Input Characteristic:-

The input characteristics is a plot of Input voltage v/s Input current keeping output voltage constant .

ie. V_{BE} v/s I_B keeping V_{CE} constant

The input characteristics is obtained by varying V_{BE} in steps and noting down the corresponding values of I_B keeping V_{CE} constant. A family of curves can be obtained for different values of V_{CE} .

The dynamic input resistance r_i is obtained using the relation:

$$r_i = \frac{\partial V_{BE}}{\partial I_B} \mid V_{CE} = \text{constant}$$

Output Characteristic:-

The output characteristic is a plot of output voltage v/s output current keeping input current constant, ie. V_{CE} v/s I_C keeping I_B constant.

It is obtained by varying V_{CE} in steps and noting down the corresponding values of I_C keeping I_B constant.

A family of curves is obtained for different values of I_B . The slope of the output characteristic is appreciable, this means that the output resistance is not as large as in case of common-base configuration. The dynamic output resistance is given by the relation:

$$r_o = \frac{\partial V_{CE}}{\partial I_C} \mid I_B = \text{constant}$$

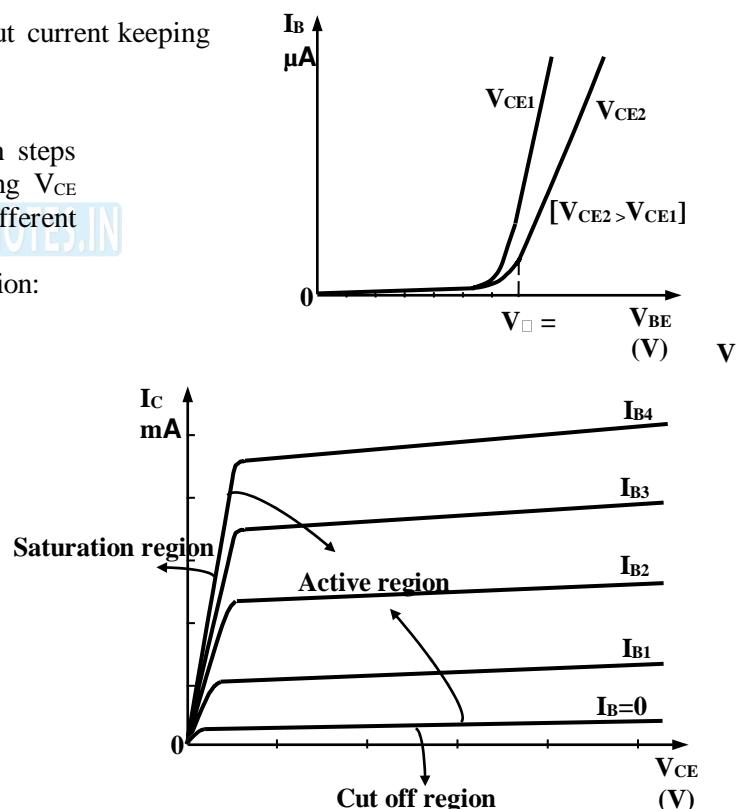
\square is the current gain of a transistor in common-emitter configuration and is determined using the relation:

$$\square = \frac{\partial I_C}{\partial I_B} \mid V_{CE} = \text{constant}$$

The input resistance r_i is high when compared to common-base configuration (500 Ω to 1.5k Ω).The output resistance has a high value (15 k Ω to 50 k Ω). The current gain β has a very large value (200 to 400). The voltage gain is very high (250 to 500).

Applications:- A transistor in common-emitter configuration is used as a voltage amplifier, power amplifier and multi-stage amplifier.

Figure shows the circuit diagram for a transistor in common-emitter configuration. Battery V_{BB} is used to forward bias the base-emitter junction. The microammeter measures the input current I_B and the voltmeter measures the input voltage V_{BE} . Battery V_{CC} is used to reverse bias the collector-base junction ($V_{CC} > V_{BB}$). The milliammeter is used to measure the output current I_C while the voltmeter measures the output voltage V_{CE} . R_B is the base resistor and R_C is the collector resistor. Here the emitter lead is common to both the input and output circuits, hence it is known as common-emitter configuration



Important Equations For Transistor In Common Emitter Configuration:-

- | | |
|---|---|
| [1] $I_C = \beta \cdot I_B + (1+\beta) I_{CEO} = \beta \cdot I_B + I_{CEO}$ | ; [2] $I_C = \beta \cdot I_B$ (neglecting leakage current, ie. $I_{CEO} = 0$). |
| [3] $I_C = I_{CEO}$ (when input current $I_B = 0$) | ; [4] $I_{CEO} = (1 + \beta) I_{CBO}$. |

COMMON COLLECTOR CONFIGURATION :-

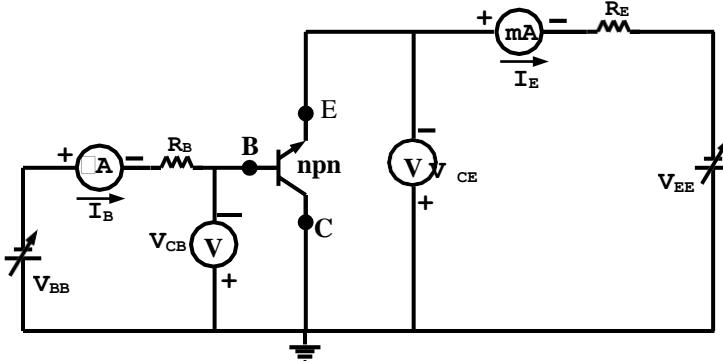


Figure shows the circuit diagram for a npn transistor in common-collector configuration. Battery V_{BB} is used to reverse bias the collector-base junction. The microammeter measures input current I_B and the voltmeter measures input voltage V_{CB} . Battery V_{EE} along with V_{BB} is used to forward bias the base-emitter junction (V_{EE} is at lower potential than V_{BB}). The milliammeter is used to measure output current I_E , while the voltmeter measures the output voltage V_{CE} . (I_E is the output current and V_{CE} is the output voltage). R_B is the base resistor and R_E is the emitter resistor. Here the collector lead is common to both the input and output circuits, hence it is known as common-collector configuration.

Input Characteristic:-

The input characteristics is a plot of Input voltage v/s Input current keeping output voltage constant .

i.e. V_{CB} v/s I_B keeping V_{CE} constant.

The input characteristics is obtained by varying V_{CB} in steps and noting down the corresponding values of I_B keeping V_{CE} constant. A family of curves can be obtained for different values of V_{CE} .

The dynamic input resistance r_i is obtained using the relation:

$$r_i = \frac{\partial V_{CB}}{\partial I_B} \mid V_{CE} = \text{constant.}$$

Output Characteristic:-

The output characteristic is a plot of Output voltage v/s Output current keeping Input current constant,

i.e. V_{CE} v/s I_E Keeping I_B constant.

It is obtained by varying V_{CE} in steps and noting down the

corresponding values of I_E keeping I_B constant.

A family of curves are obtained for different values of I_B .

$I_C \approx I_E$ because I_B is very small.

The slope of the output characteristic is more appreciable than that of the common-emitter configuration, this means that the output resistance is much smaller than that of the common-emitter configuration. The dynamic output resistance is given by the relation:

$$r_o = \frac{\partial V_{CE}}{\partial I_E} \mid I_B = \text{constant.}$$

γ is the current gain of a transistor in common-collector configuration and is determined using the relation:

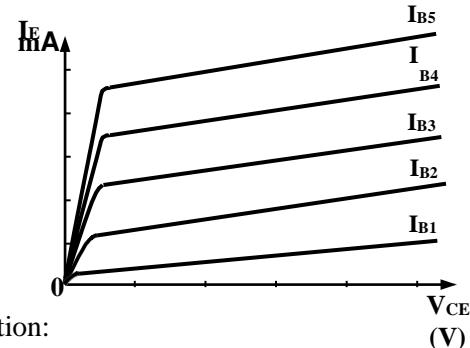
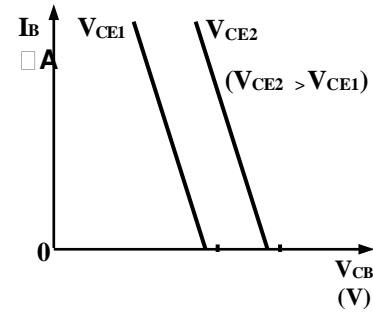
$$\gamma = \frac{\partial I_E}{\partial I_B} \mid V_{CE} = \text{constant.}$$

The input resistance r_i is very high when compared to common-emitter configuration ($\approx 1 \text{ M}\Omega$). The output resistance has a very low value (less than 500Ω). The current gain γ has a very large value ($\beta + 1$). The voltage gain is very low ($A_v \approx 1$).

Applications:- It is used as a buffer amplifier to provide excellent impedance matching between two stages. The circuit is commonly known as an **Emitter follower**.

Comparison Of The Three Transistor Configurations:-

Sl. No.	PARAMETER	C-B Confgrn.	C-E Confgrn.	C-C Confgrn.
1	Current gain	Very low (<1)	Very high (200-400)	Very high ($1+\beta$)
2	Voltage gain	High (100-200)	Very high (250-500)	Very low (<1)
3	Input impedance	Very low ($10-15\Omega$)	Medium ($\sim 1 \text{ k}\Omega$)	Very high ($\sim 1 \text{ M}\Omega$)
4	Output impedance	Very high ($\approx 1 \text{ M}\Omega$)	Medium (20-50 $\text{k}\Omega$)	Very low (< 1 $\text{k}\Omega$)
5	Applications	Wide-band Amp, Constant current source, Buffer Amp.	Voltage Amp, multi-stage Amp, power-Amp, Audio Amp.	Buffer Amp. (as Impedance matching network)



RELATIONSHIP BETWEEN α AND β :

- α is the current gain of a transistor in common-base configuration and is given by the relation, $\alpha = \frac{\Delta I_C}{\Delta I_B}$
- β is the current gain of a transistor in common-emitter configuration and is given by the relation, $\beta = \frac{\Delta I_C}{\Delta I_E}$

- (1) α in terms of β :- The basic transistor equation is given by: $I_E = I_B + I_C$.----- (A)
 Considering the incremental values, we have $\Delta I_E = \Delta I_B + \Delta I_C$.----- (B)

Divide equation (B) through out by ΔI_C , ie. $(\Delta I_E / \Delta I_C) = (\Delta I_B / \Delta I_C) + (\Delta I_C / \Delta I_C)$ ----- (C)

But $(\Delta I_C / \Delta I_E) = \alpha \Rightarrow (\Delta I_E / \Delta I_C) = (1 / \alpha)$ & $(\Delta I_C / \Delta I_B) = \beta \quad (\Delta I_B / \Delta I_C) = (1 / \beta)$

ie. $(1 / \alpha) = (1 / \beta) + 1$ ----- (D).

ie. $(1 / \alpha) = (1 + \beta) / \beta$; Taking the reciprocal we have :

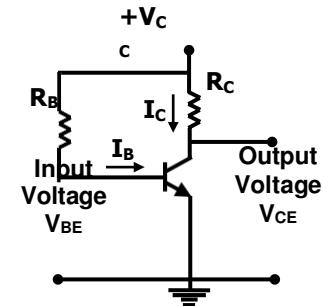
$$\alpha = \beta / (\beta + 1)$$

- (2) β in terms of α :- From equation (D), we have $1 / \alpha = 1 + (1 / \beta) \quad 1 / \beta = (1 / \alpha) - 1$ ie.

$$(1 / \beta) = (1 - \alpha) / \alpha \quad \beta = \alpha / (1 - \alpha)$$

FIXED BIAS CIRCUIT or BASE BIAS CIRCUIT : Figure shows a fixed bias circuit . V_{CC} is the battery used for biasing both the junctions. R_C is the collector resistance & R_B is the base resistance. The Q-point is located in the active region by properly selecting the values of V_{CC} , R_B & R_C , so that a proper value of base current (I_B) will fix up the quiescent I_{CQ} & V_{CEQ} & hence the operating point.

The operating point position on the load line can be determined by calculating the values of I_C & V_{CE} in the circuit.



(1) The output Current or Collector Current (I_C):

The output current or the Collector current is given by the expression :

$$I_C = \beta \cdot I_B + I_{CEO} \quad [\text{but } I_{CEO} \text{ is very small compared to } \beta \cdot I_B]$$

Hence $I_C = \beta \cdot I_B$ ----- (A)

But the base current or input current is given by : $I_B = [V_{CC}/R_B]$

Therefore the collector current is given by : $I_C = \beta \cdot [V_{CC}/R_B]$ ----- (B)

(2) The output voltage or the Collector- Emitter Voltage (V_{CE}):

The out put voltage or the Collector-Emitter voltage is given by:

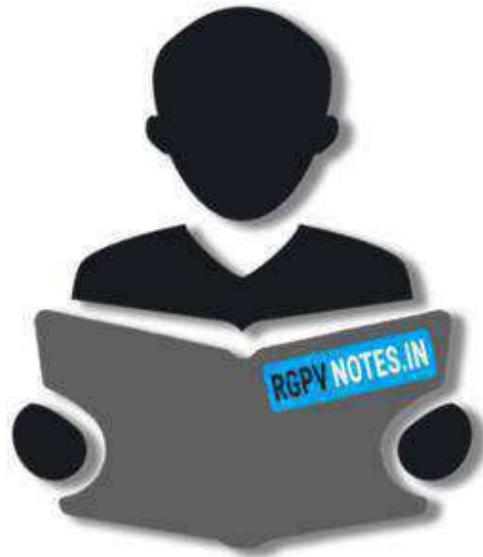
$$V_{CC} = I_C \cdot R_C + V_{CE}$$

$$\text{ie. } V_{CE} = [V_{CC} - I_C \cdot R_C]$$
----- (C)

Use the value of I_C from equation (B) in equation (C) ,

$$\text{ie. } V_{CE} = V_{CC} [1 - \beta \cdot (R_C/R_B)]$$
----- (D)

The values of I_C and V_{CE} obtained using equations (B) and (D) will help in locating the operating point of the transistor on the DC load line .



RGPVNOTES.IN

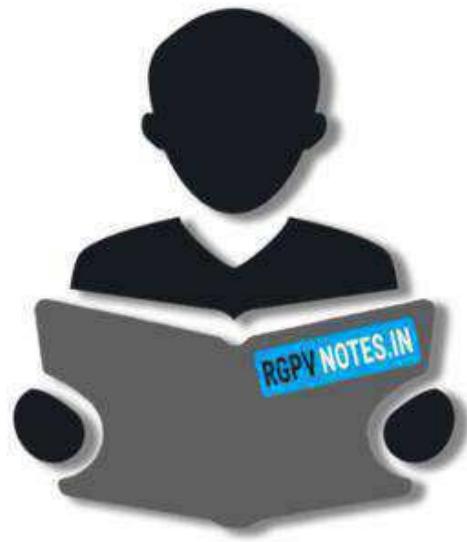
We hope you find these notes useful.

You can get previous year question papers at
<https://qp.rgpvnotes.in> .

If you have any queries or you want to submit your
study notes please write us at
rgpvnotes.in@gmail.com



LIKE & FOLLOW US ON FACEBOOK
facebook.com/rgpvnotes.in



RGPVNOTES.IN

Program : **B.Tech**

Subject Name: **Engineering Chemistry**

Subject Code: **BT-101**

Semester: **1st**



LIKE & FOLLOW US ON FACEBOOK
facebook.com/rgpvnotes.in

UNIT V

Phase equilibrium and corrosion

Phase- It is defined as any homogeneous, physically distinct and mechanically separable portion of a system, which is separated from other such parts of the system by definite boundary surfaces in a system.

Component- It is defined as the smallest number of independently variable constituents taking part in the state of equilibrium by means of which the composition of each phase can be expressed directly or in the form of chemical equation.

Degree of freedom- It is defined as the minimum number of the independently variable factors such as the temperature, pressure and composition of the phases which must be arbitrarily specified in order to represent perfectly the condition of a system.

PHASE RULE

Gibbs' phase rule was proposed by Josiah Willard Gibbs . The rule applies to non-reactive multi-component heterogeneous systems in thermodynamic equilibrium and is given by the equality .

$$F = C - P + 2$$

where F is the number of degrees of freedom, C is the number of components and P is the number of phases in thermodynamic equilibrium with each other.

The number of degrees of freedom is the number of independent intensive variables, i.e. the largest number of thermodynamic parameters such as temperature or pressure that can be varied simultaneously and arbitrarily without determining one another. An example of one-component system is a system involving one pure chemical, while two-component systems, such as mixtures of water and ethanol, have two chemically independent components, and so on. Typical phases are solids, liquids and gases.

Uses of Phase Rule : -

- With the help of terms phases , components and degrees of freedom the phase rule gives a convenient basis for classification of equilibrium state of system.
- The information about molecular structure is not necessary because It applies to macroscopic systems .
- The phase rule shows that various systems having same degrees of freedom behave in a similar fashion .
- It helps in deciding whether the given number of substances together would exist in equilibrium under a given set of conditions

Limitations of phase rule : -

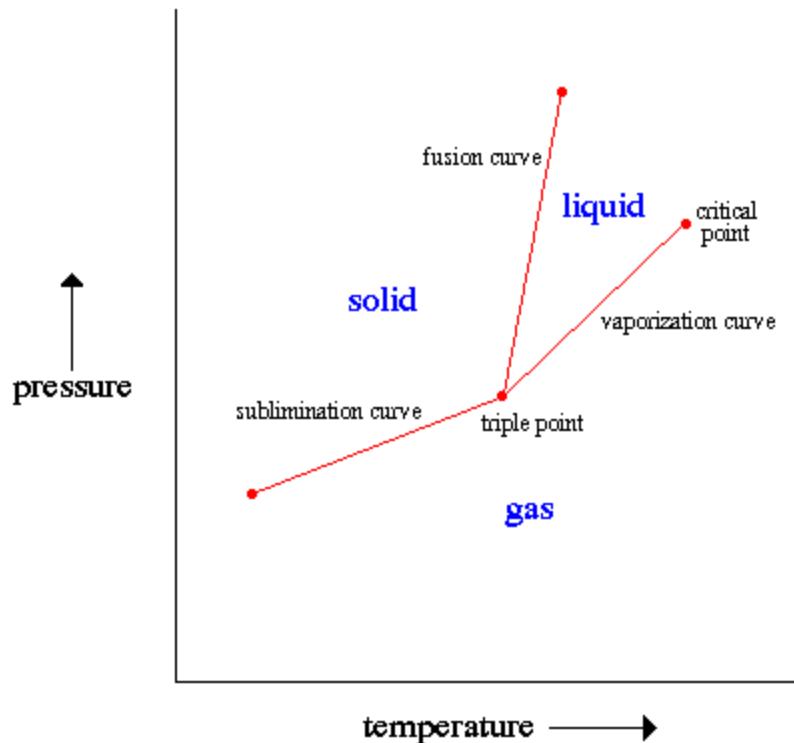
- Not applicable for the systems which are slow in reaching to equilibrium state .
- There is no consideration of magnetic and electric influence since system is defined only the variables like pressure temperature and components force.

- Any liquid or solid present should not be in finally divided state otherwise the value of their vapor pressure will differ from their normal value .

Phase equilibrium – it refers to the set of conditions where more than one phase may exist. It can be reflected by constancy with time in the phase characteristics of a system. In most metallurgical and materials systems, phase equilibrium involves just solid phases. However the state of equilibrium is never completely achieved because of very slow rate of approach of equilibrium in solid systems. This leads to non-equilibrium or meta-stable state, which may persist indefinitely and of course, has more practical significance than equilibrium phases. An equilibrium state of solid system can be reflected in terms of characteristics of the microstructure, phases present and their compositions, relative phase amounts and their spatial arrangement or distribution.

A **phase diagram** in physical chemistry, engineering, mineralogy, and materials science is a type of chart used to show conditions (pressure, temperature, volume, etc.) at which thermodynamically distinct **phases** occur and coexist at equilibrium. Common components of a phase diagram are lines of equilibrium or phase boundaries, which refer to lines that mark conditions under which multiple phases can coexist at equilibrium. Phase transitions occur along lines of equilibrium.

A phase diagram is a graph showing the limiting conditions for solid, liquid, and gaseous phases of a single substance or of a mixture of substances while undergoing changes in pressure and temperature or in some other combination of variables, such as solubility and temperature. The figure shown below displays a typical phase diagram for a one-component system (i.e., one consisting of a single pure substance), the curves having been obtained from measurements made at various pressures and temperatures.



Phase diagrams are specific for each substance and mixture. Complex mixtures may require three-dimensional phase diagrams, which can be represented in two dimensions through use of perspective. Phase diagrams are widely used in studies of mineral equilibriums in connection with the conditions of formation of rocks and minerals within the Earth. They also are invaluable when designing industrial equipment and seeking optimum conditions for manufacturing processes, and in determining the purity of substances.

Phase diagrams illustrate the variations between the states of matter of elements or compounds as they relate to pressure and temperatures. The following is an example of a phase diagram for a generic single-component system:

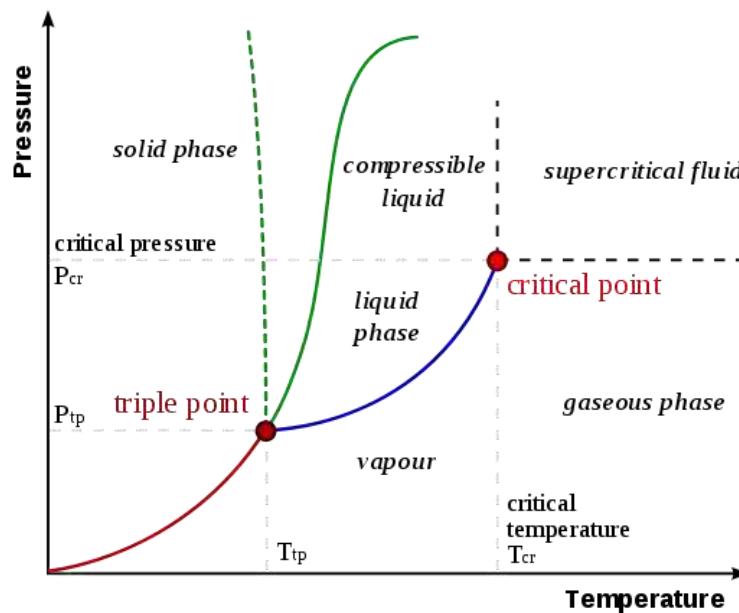


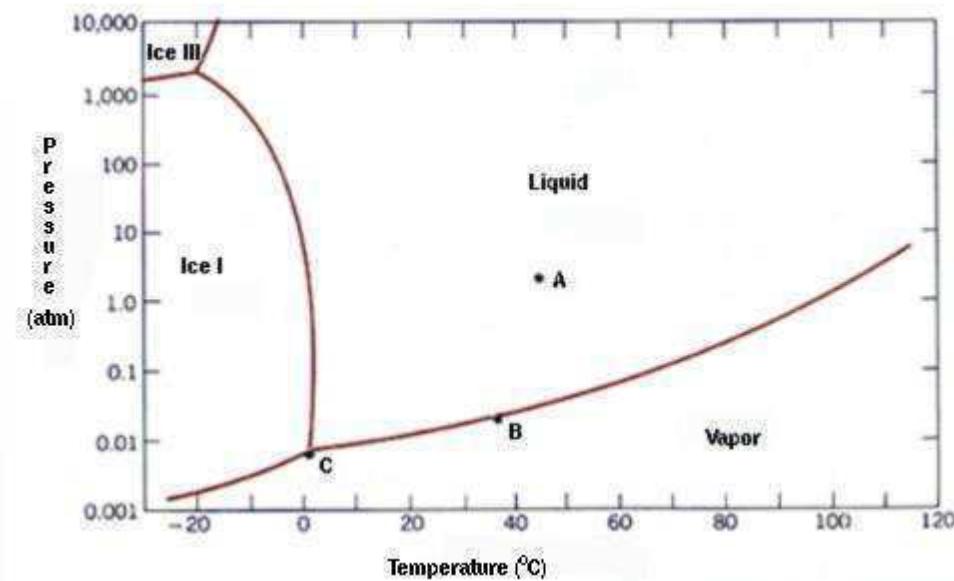
Figure 1. General Phase diagram

- **Triple point** – the point on a phase diagram at which the three states of matter: gas, liquid, and solid coexist
- **Critical point** – the point on a phase diagram at which the substance is indistinguishable between liquid and gaseous states
- **Fusion(melting) (or freezing) curve** – the curve on a phase diagram which represents the transition between liquid and solid states
- **Vaporization (or condensation) curve** – the curve on a phase diagram which represents the transition between gaseous and liquid states
- **Sublimation (or deposition) curve** – the curve on a phase diagram which represents the transition between gaseous and solid states

Binary diagrams: These diagrams constitute two components, e.g.: two metals (Cu and Ni), or a metal and a compound (Fe and Fe₃C), or two compounds (Al₂O₃ and Si₂O₃), etc. In most engineering applications, as mentioned before, condensed phase rule is applicable. It is assumed that the same is applicable for all binary diagrams, thus the presentation of binary diagrams becomes less complicated. Thus binary diagrams are usually drawn showing variations in temperature and composition only. It is also to be noted that all binary systems consist only one liquid phase i.e. a component is completely soluble in the other component when both are in liquid state. Hence, binary systems are classified according to their solid solubility. If both the components are completely soluble in each other, the system is called isomorphous system. E.g.: Cu-Ni, Ag-Au, Ge-Si, Al₂O₃-Cr₂O₃. Extent solid solubility for a system of two metallic components can be predicted based on Hume-Ruthery conditions, summarized in the following: - Crystal structure of each element of solid solution must be the same. - Size of atoms of each two elements must not differ by more than 15%. - Elements should not form compounds with

each other i.e. there should be no appreciable difference in the electro-negativities of the two elements. - Elements should have the same valence.

Isomorphous system: Figure-2 depicts a typical phase diagram for an isomorphous system made of two metallic elements A and B. As cited earlier, any phase diagram can be considered as a map. A set of coordinates – a temperature and a composition – is associated with each point in the diagram. If the alloy composition and temperature specified, then the phase diagram allows determination of the phase or phases that will present under equilibrium conditions. There are only two phases in the phase diagram, the liquid and the solid phases. These single-phases regions are separated by a two-phase region where both liquid and solid co-exist. The area in the figure-2 above the line marked liquids ($A'bB'$) corresponds to the region of stability of the liquid phase, and the area below the solidus line ($A'dB'$) represents the stable region for the solid phase.



CORROSION

Corrosion is defined as “the destruction or deterioration and consequent loss of metals or alloys through chemical or electrochemical attack by the surrounding environment”.

The primary factors that initiate corrosion on metals are atmospheric air, water and also conducting surface of the metal.

Eg: Rusting of iron, green scales are formed on copper vessels

Corrosion of metal occurs either by direct chemical attacks or by electrochemical attack on the metal by the corrosive environment.

If the corrosion takes place due to direct chemical attack (in the absence of moisture) that type of corrosion is known as dry corrosion.

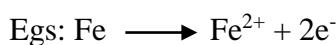
If the corrosion of metal takes place due to electrochemical attack in presence of moisture or a conducting medium such corrosion is known as wet corrosion or electrochemical corrosion.

Electrochemical theory of corrosion: Most of the corrosion takes place on the basis of electrochemical reactions on the surface of metal such a type of corrosion is known as wet corrosion. Electrochemical theory of corrosion can be taking iron as an example.

When a metal like iron is exposed to the environment according to electrochemical theory corrosion of metal takes place due to the formation of anodic and cathodic regions on the same metal surface or when the two metals are in contact with each other in a corrosive medium. These anodes and cathodes are formed due to the heterogeneities at the interfaces of the metal and environment. The heterogeneities on a metal surface could develop due to several factors like

1. On a metal surface if the concentration of the oxygen is different (if in the metal the area which is exposed to more oxygen acts as cathode, the area which is exposed to less oxygen concentration acts as anode).
2. Due to contact of two different metals (egs: if copper and iron are in contact with each other, then Fe acts as anode and copper acts as cathode due to change in electrode potential).
3. If metal surface subjected to stress (area under stress acts as anode).

Thus anodic and cathodic area are formed, in presence of corrosion medium (like moisture etc.) At anode oxidation takes place so that metal is converted into metal ions with the liberation of electrons.

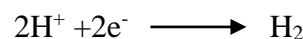


At the cathodic regions, reduction takes place since the metal at cathodic region cannot be reduced further, so some constituents of the corrosive medium take part in the cathodic reaction. Since in the cathodic reaction as the constituents of the corrosion medium are involved, they are more complicated and dependent on the nature of environment. Most common type of cathodic reaction are 1. Liberation of hydrogen 2. Absorption of oxygen.

Cathodic reactions:

(a) Liberation of Hydrogen type

- (i) If the medium is acidic and in the absence of O₂

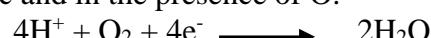


- (ii) If the medium is neutral or alkaline in the absence of O₂.

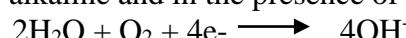


(b) Absorption of oxygen type

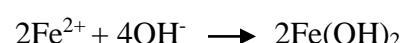
- (i) If the medium is acidic and in the presence of O₂.



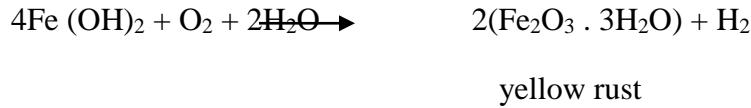
- (ii) If the medium is neutral or alkaline and in the presence of O₂.



The metal ions (Fe²⁺) liberated at anode and some anions (OH⁻) formed at cathode diffuse towards each other through the conducting medium and form a corrosion product somewhere between the anode and cathode as



In an oxidizing environment, the insoluble Fe(OH)_2 oxidised to ferric oxide as following reaction.



if the concentration of oxygen is limited then Fe(OH)_2 is converted into magnetic oxide of Fe and is known as black rust.



Mechanism of CORROSION : - The anodic reaction would be



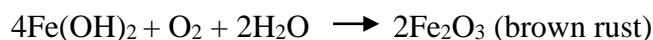
The loss of electrons leaves positively charged ions at the anode which travel from the anode to cathode via the water and carry the positive current. The electrons which are released on the anode travel from the anode to the cathode via the metallic circuit. These electrons are utilized in the reduction of oxygen present in water which is in contact with the cathode. Hence, at the cathode, one of the following reactions may take place.



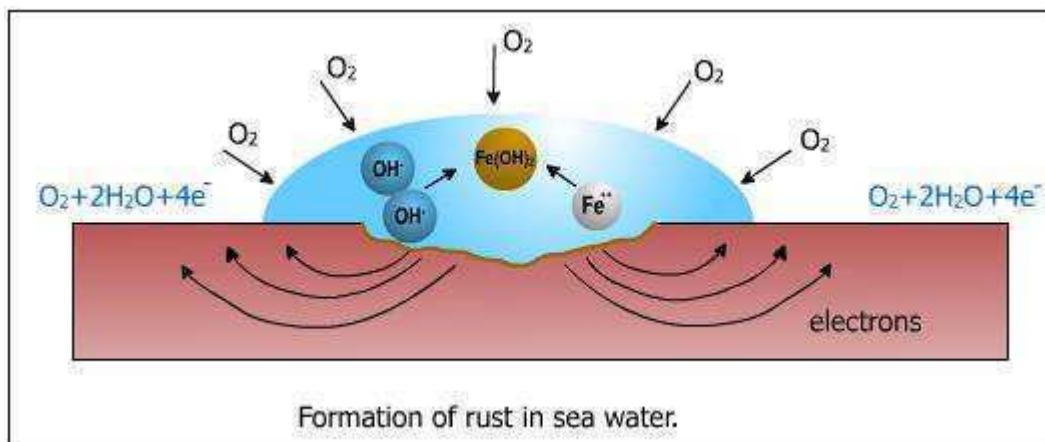
The negatively charged OH^- ions react with the positively charged Fe^{2+} ions and forms Fe(OH)_2 .



OR



The end product is Fe_2O_3 , which is generally called "rust". The rust is formed a little away from the surface. The mechanism of rust formation is shown in the following figure.



Types of Corrosion: Corrosion on the metals taking place depending on the nature of metals and depending on the types of environment by different mechanisms, giving different types of corrosion.

1 Galvanic corrosion or differential metal corrosion:

This occurs when two dissimilar metals are in contact with each other in a corrosive conductive medium; a potential difference is set up resulting in a galvanic current. The two metals differ in their tendencies to undergo oxidation. The metal with lower electrode potential or more active metal acts as anode and the metal with higher electrode potential acts as cathode. The potential difference is main factor for corrosion to take place. The anodic metal undergoes corrosion whereas cathodic metal gets un -attacked.

Ex: When iron contact with copper iron has lower electrode potential acts as anode and undergoes oxidation as,



Whereas copper which is having higher electrode potential acts as cathode gets unaffected. The rate of galvanic corrosion depends upon potential difference between anodic and cathodic

metals, ratio of anodic and cathodic area and environmental factors and tendency of the metal to undergo passivity etc.

Other example : When Fe contact with Sn then Fe acts as anode and Sn acts as cathode but when Fe contact with Zn, Fe acts as cathode where as Zn acts as anode.

PITTING CORROSION: Pitting corrosion is a localized and accelerated corrosion. When a small particles of dust or water etc are get deposited on a metal (like steel). The portion covered by the dust will not be well-aerated area compared to the exposed surface hence the covered surface becomes anodic with respect to the surface exposed. In presence of an conducting medium (moisture) corrosion starts below the dust part and forming a pit. Once pit is formed the ratio of corrosion increases, because of the formation of smaller anodic and larger cathodic area intense corrosion takes place.

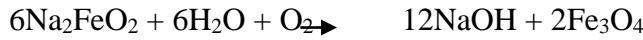
Pitting corrosion is one of the most destructive forms of corrosion. It causes equipment to fail because of perforation with only a small percent weight loss of the entire structure. Because of the small sizes of the pits it is highly difficult to identify the pitting corrosion. Pitting corrosion is an autocatalytic process, and once the corrosion products are formed, it further provides the condition for differential aeration below the corrosion product and surrounding metal parts. The pit grows and ultimately may cause failure of the metal.

Caustic embrittlement: It is a form of stress corrosion takes place in boilers operating at high temperature and pressure. Caustic embrittlement focus at stressed part of boilers such as cracks, rivets, bents, joints etc.

The boiler fed water usually contains some residual sodium carbonate (used for softening process). At high temperature and pressure it undergoes hydrolysis to form sodium hydroxide.



The alkali water sweeps through the minute cracks, crevices between the rivets and joints by capillary action. Inside the cracks water gets evaporated leaving behind NaOH. The concentrations of the NaOH gradually increase on these sites due to poor circulation of water. When concentrations of the NaOH reaches a value of 10% it attacks the metal at the stressed region dissolving it in the form of sodium ferroate (Na_2FeO_2). Sodium ferroate undergoes hydrolysis-depositing magnetite as follows



So NaOH is regenerated in the process and its concentration is keep on increasing maintaining a required environment. Thus corrosion process develops cracks and making the metal brittle by the deposition of the product.

The corrosion cell can be represented as

Fe (under stress) / conc. NaOH / dil. NaOH/ Fe (stress free)

Caustic embrittlement can be prevented by the addition of compounds like sodium sulphite, tannin, lignin, phosphates etc. which blocks the cracks thereby preventing the infiltration of alkali.

Factors affecting rate of corrosion: Several factors affecting the rate of the corrosion, which can be divided into two parts

2. Factors affecting on metal (related to the metal)
 3. Factors affecting on corrosive environment.

Factors affecting the metals:

- 1. Nature of the metal:** The tendency of the metal to undergo corrosion is mainly dependent on the nature of the metal. In general the metal with lower electrode potential have more reactive and more susceptible for corrosion and metal with high electrode potential are less reactive and less susceptible for corrosion for egs: metals like K, Na, Mg, Zn etc have low electrode potential are undergo corrosion very easily, where as noble metals like Ag, Au, Pt have higher electrode potential, their corrosion rate are negligible but there are few exception for this general trend as some metals show the property of passivity like Al, Cr, Ti, Ta etc.
 - 2. Surface state of the metal or nature of the corrosion product (passivity):** The corrosion product is usually the oxide of the metal; the nature of the product determines the rate of further corrosion process.

If the oxide layer, which forms on the surface, is stoichiometric, highly insoluble and non-porous in nature with low ionic and electronic conductivity then that type of products

layer effectively prevents further corrosion, which acts as a protective film. For egs: Al, Cr, Ti develop such a layer on their surface and become passive to corrosion and some metal like Ta, Zr and Mo not only forms such a protective layers but are capable of self repairing oxide films when it is damaged. Hence these are extremely passive metals.

If the oxide layer forms on the metal surface is non-stoichiometric, soluble, unstable and porous in nature and have an appreciable conductivity, they cannot control corrosion on the metal surface for egs: oxide layer formed on metals like Zn, Fe, Mg etc.

3. Anodic and Cathodic area:

The rate of the corrosion is greatly influenced by the relative sizes of cathodic and anodic areas.

If the metal has smaller the anodic area and larger the cathodic area exposed to corrosive atmosphere, more intense and faster is the corrosion occurring at anodic area because at anode oxidation takes place and electrons are liberated. At the cathode these electrons are consumed. When anode is smaller and cathode region is larger all the liberated electrons at anode are rapidly consumed. This process makes the anodic reaction to takes place at its maximum rate thus increasing the corrosion rate. If the cathode is smaller and reverse process takes place decrease rate of corrosion.

For egs: If tin (Sn) coated on iron (Fe) and in that some area are not covered or some pin holes are left, there forms smaller anodic area and larger cathodic area because tin is cathodic with respect to iron so intense localized corrosion takes place. On the other hand if Zn coated to Fe then if there are some pin holes are there creates larger anodic area and smaller cathodic area because Fe is cathodic with respect to zinc so that rate of corrosion is very less.

4. Hydrogen over voltage:

A metal with low hydrogen over voltage on its surface is more susceptible for corrosion. When the cathodic reaction is hydrogen evolution type with low hydrogen over voltage, liberation of H₂ gas is easier so that cathodic reaction is very fast, that makes anodic reaction faster hence overall corrosion process is very fast. If the H₂ over voltage is high so cathodic reaction is slow hence corrosion reaction also slower.

Factors related to corrosive environment:

1. pH of the medium: Usually higher acidic nature (low pH) higher is the rate of corrosion. If the pH is greater than 10 corrosion of iron is very less due to the formation of protective coating of hydrous oxides of iron.

If pH is between 10 and 3, then presence of oxygen is essential for corrosion of iron. If the pH is 3 or lower than 3, severe corrosion occurs in the absence of air due to the continuous evolution of H₂ at cathode. However metals like Al, Zn etc undergo fast corrosion in highly alkaline medium.

2. Temperature: On increasing the temp. Rate of corrosion process also gets increases because on increase of temp. Conductance of the aqueous medium increases hence rate of diffusion also.

In some cases on rise in temp. Decrease the passivity, which again leads to increase in the corrosion rate.

3. Polarization at anodic and cathodic area: Polarization of cathode or anode decreases the rate of corrosion. If anodic polarization takes place due to some reaction, then tendency of metal to undergo oxidation decreases hence dissolution of metals as metal ion decreases. This is usually due to increase in conc. Of ions of the dissolved metals in the vicinity of electrode or also due to the anodic passivity. Cathode polarization decreases the cathodic reaction hence hindering the combination of cathode reactant and electron. For the corrosion to continue both anodic and cathodic reaction should take place simultaneously if any one reaction is slower than the rate of corrosion is slower. Use of depolarizers reduces the polarization effect hence the rate of corrosion reaction increases.

Galvanic corrosion or Corrosion of Metals due to contact of metals:

The name of this type of corrosion is indicative that there must be a formation of galvanic cell on metal surface causing corrosion. Such cells get set up all along the surface of metal when it faces the electrolytic environment and two dissimilar metals in contact.

For example:

- If metal iron or brass alloy is used together to join various parts, due to difference in potential, a galvanic cells gets set, leading to corrosion.
- In such cases all along the surface of metal small galvanic cells are set up, where area of high potential acts as an anode and the one with lower potential act as a cathode. The portion of metal acting as an anode deteriorates and at cathode the cathodic product gets evolved/deposited, depending upon the environment, and mechanism of the corrosion, i.e. by absorption of oxygen or evolution of hydrogen.
- When metals are surrounded by aqueous alkaline/acidic/neutral solution, the movement electrons, cations and anions occurs just like as in a typical galvanic cell. But when the

metal is just exposed to moisture or a few drops of certain solution, the movement of electrons, cations and anions takes place through the bulk of metal, thus the metal itself serves as a conducting medium while its outer surface gets corroded.

- If the metal is exposed to acidic medium the mechanism at cathode proceeds to form H₂ gas while anodic reactions remain the same.

Control of corrosion: *To avoid the formation of galvanic cells,*

- (a) The metals should be pure.
- (b) The materials used to assemble the different parts should be of same potentials.



RGPVNOTES.IN

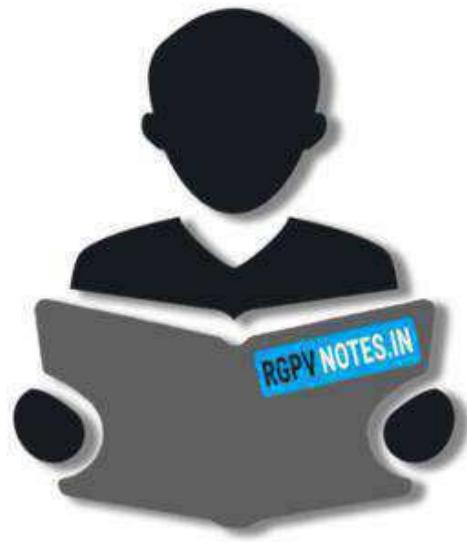
We hope you find these notes useful.

You can get previous year question papers at
<https://qp.rgpvnotes.in> .

If you have any queries or you want to submit your
study notes please write us at
rgpvnotes.in@gmail.com



LIKE & FOLLOW US ON FACEBOOK
facebook.com/rgpvnotes.in



RGPVNOTES.IN

Program : **B.Tech**

Subject Name: **Engineering Graphics**

Subject Code: **BT-105**

Semester: **1st**



LIKE & FOLLOW US ON FACEBOOK
facebook.com/rgpvnotes.in

Module V

Isometric Projections covering, Principles of Isometric projection – Isometric Scale, Isometric Views, Conventions; Isometric Views of lines, Planes, Simple and compound Solids

Orthographic Projections:**Introduction**

Any object has three dimensions, viz., length, width and thickness. A projection is defined as a representation of an object on a two dimensional plane. The projections of an object should convey all the three dimensions, along with other details of the object on a sheet of paper. The elements to be considered while obtaining a projection are:

- (i) The object
- (ii) The plane of projection
- (iii) The point of sight
- (iv) The rays of sight

A projection may be obtained by viewing the object from the point of sight and tracing in correct sequence, the points of intersection between the rays of sight and the plane on to which the object is projected. A projection is called orthographic projection when the point of sight is imagined to be located at infinity so that the rays of sight are parallel to each other and intersect the plane of projection at right angle to it.

The principles of orthographic projection may be followed in four different angles or systems, viz., first, second, third and fourth angle projections. A projection is said to be first, second, third or fourth angle when the object is imagined to be in the first, second, third or fourth quadrant respectively. However, the Bureau of Indian Standards (SP-46:1988) prefers first angle projection and throughout this book, first angle projection is followed.

Principle of First Angle Projection

In first angle projection, the object is imagined to be positioned in the first quadrant. The view from the front of the object is obtained by looking at the object from the right side of the quadrant and tracing in correct sequence, the points of intersection between the projection plane and the rays of sight extended. The object is between the observer and the plane of projection (vertical plane). Here, the object is imagined to be transparent and the projection lines are extended from various points of the object to intersect the projection plane. Hence, in first angle projection, any view is so placed that it represents the side of the object away from it.

Methods of obtaining Orthographic Views**1. View from Front**

The view from the front of an object is defined as the view that is obtained as projection on the vertical plane by looking at the object normal to its front surface. It is the usual practice to position the object such that its view from the front reveals most of the important features. Figure shows the method of obtaining the view from the front of an object.

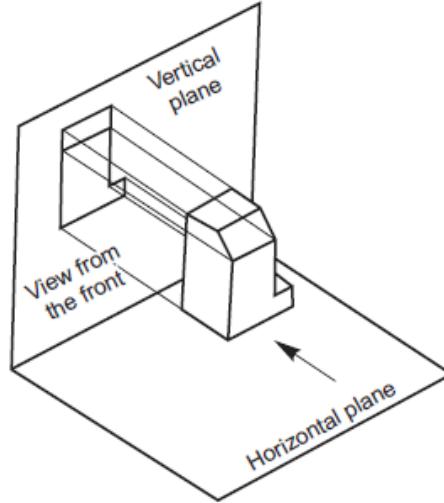


Fig. 5.1 Principle of obtaining the view from the front

2. View from above

The view from above of an object is defined as the view that is obtained as projection on the horizontal plane, by looking the object normal to its top surface. Figure shows the method of obtaining the view from above of an object.

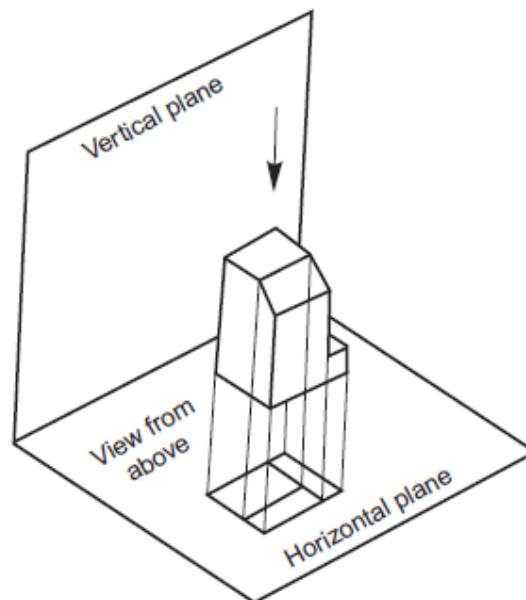


Fig. 5.2 Principle of obtaining the view from the above

3. View from the Side

The view from the side of an object is defined as the view that is obtained as projection on the profile plane by looking the object, normal to its side surface. As there are two sides for an object, viz., left side and right side, two possible views from the side, viz., view from the left and view from the right may be obtained for any object. Figure shows the method of obtaining the view from the left of an object.

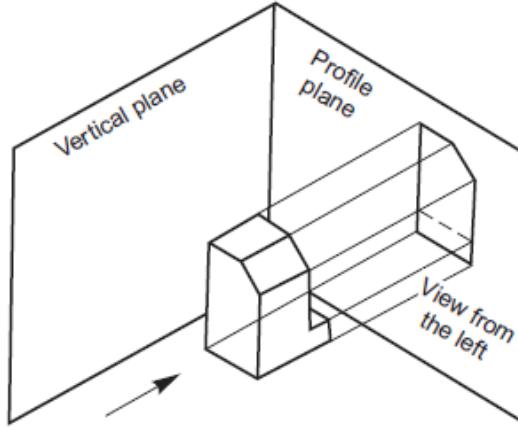


Fig. 5.3 Principle of obtaining the view from the Left

Presentation of Views

The different views of an object are placed on a drawing sheet which is a two dimensional one, to reveal all the three dimensions of the object. For this, the horizontal and profile planes are rotated till they coincide with the vertical plane. Figure shows the relative positions of the views, viz., the view from the front, above and the left of an object.

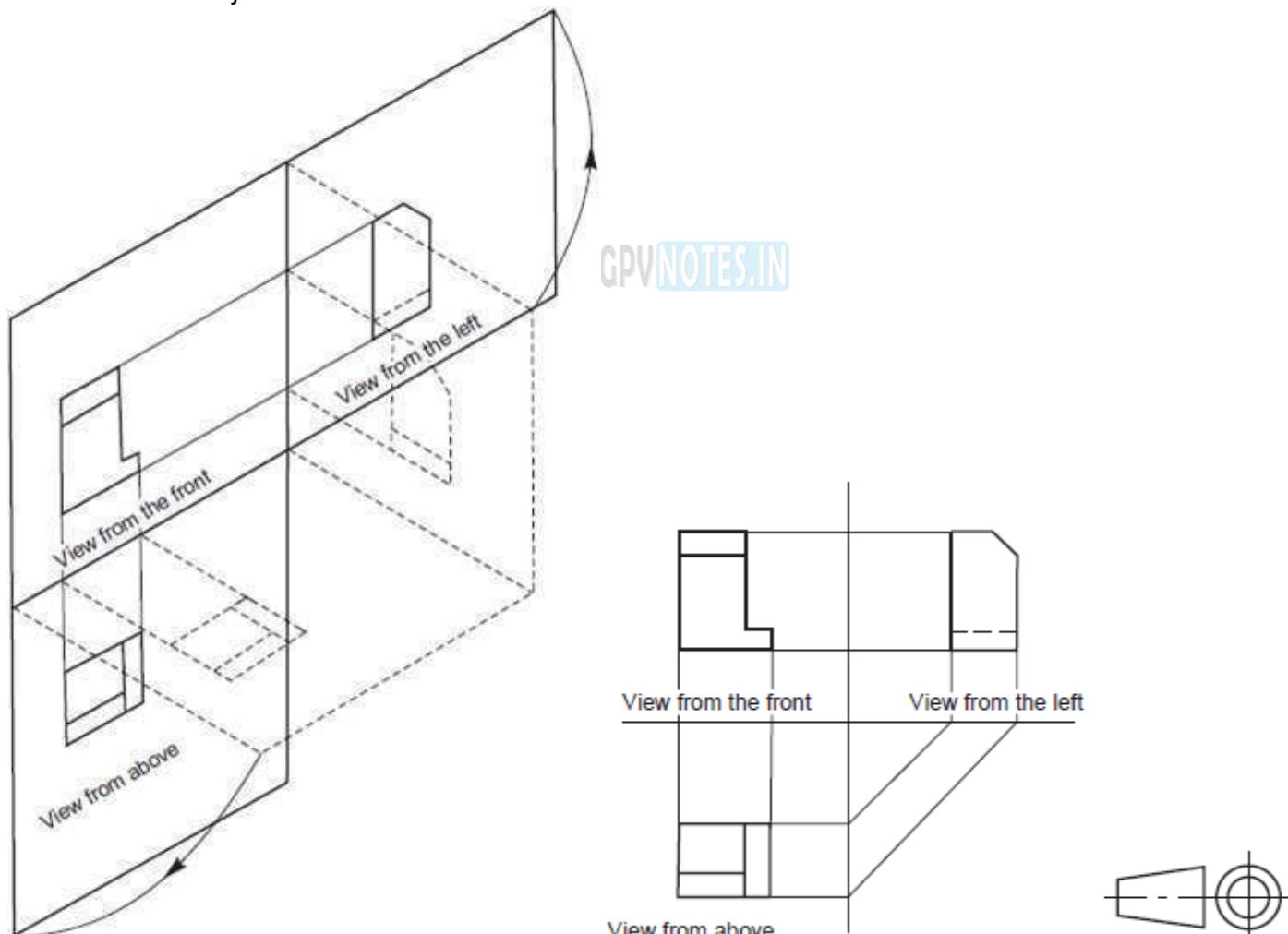


Fig. 5.5 Relative positions of the three views and the symbol

Designation and Relative Positions of Views

An object positioned in space may be imagined as surrounded by six mutually perpendicular planes. So, for any object, six different views may be obtained by viewing at it along the six directions, normal to these planes. Figure shows an object with six possible directions to obtain the different views which are designated as follows:

1. View in the direction **a** = view from the front
2. View in the direction **b** = view from above
3. View in the direction **c** = view from the left
4. View in the direction **d** = view from the right
5. View in the direction **e** = view from below
6. View in the direction **f** = view from the rear

Figure (a) shows the relative positions of the above six views in the first angle projection and Fig. (b), the distinguishing symbol of this method of projection. Figure (c) shows the relative position of the views in the third angle projection and Fig. (d), the distinguishing symbol of this method of projection.

NOTE: A comparison of Figs. (a), (b) and (c), (d) reveals that in both the methods of projection, the views are identical in shape and detail. Only their location with respect to the view from the front is different.

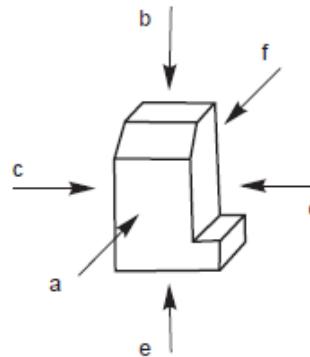


Fig. 5.6 Designation of Views

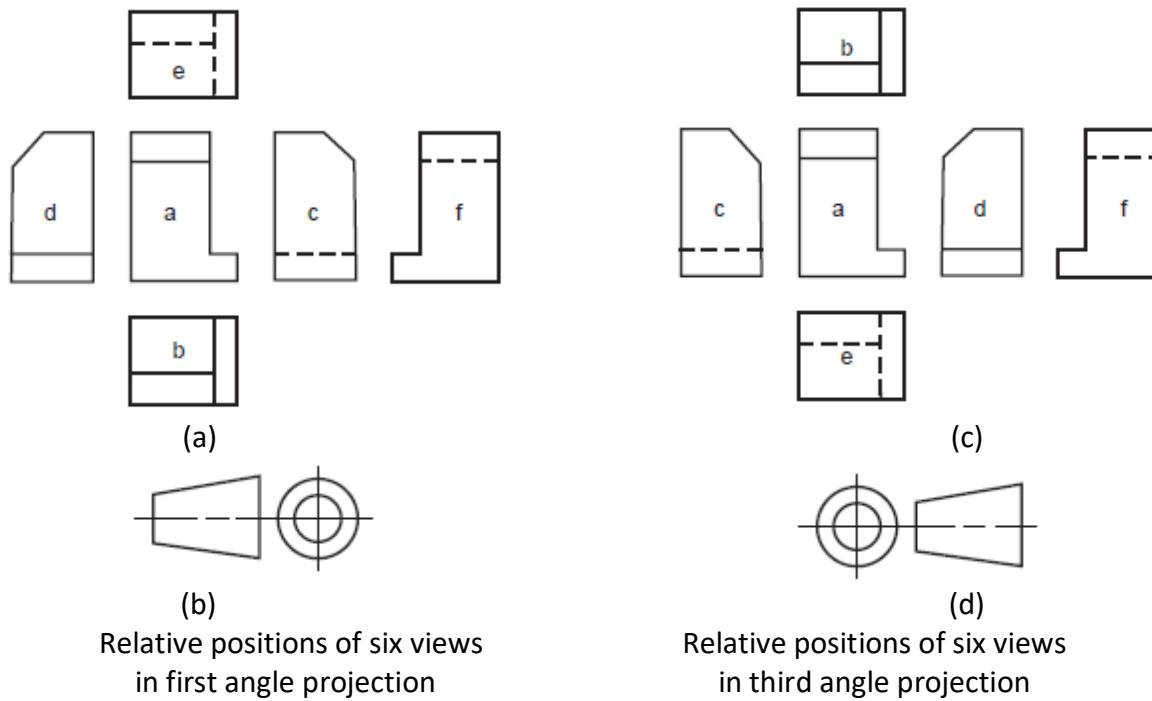


Fig. 5.7

Position of the Object

It is important to understand the significance of the position of the object relative to the planes of projection. To get useful information about the object in the orthographic projections, the object may be imagined to be positioned properly because of the following facts:

1. Any line on an object will show its true length, only when it is parallel to the plane of projection.
 2. Any surface of an object will appear in its true shape, only when it is parallel to the plane of projection.
- In the light of the above, it is necessary that the object is imagined to be positioned such that its principal surfaces are parallel to the planes of projection.

Hidden Lines

While obtaining the projection of an object on to any principal plane of projection, certain features of the object may not be visible. The invisible or hidden features are represented by short dashes of medium thickness. Figure shows the application of hidden lines in the projection of an object.

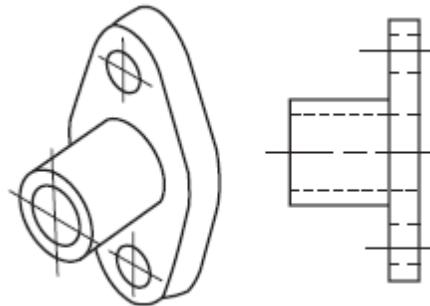


Fig. 5.8 Application of Hidden Lines

Curved Surfaces

Certain objects contain curved surfaces, tangential to other curved surfaces. The difficulty in representing the surfaces can be overcome if the following rule is observed. Wherever a tangential line drawn to the curved surface becomes a projector, a line should be drawn in the adjacent view. Figure shows the representation of certain curved surfaces, tangential to other curved surfaces.

Certain objects manufactured by casting technique, frequently contain corners filleted and the edges rounded. When the radius of a rounded corner is greater than 3 mm and the angle between the surfaces is more than 90°, no line is shown in the adjacent view. Figure shows the application of the above principle.

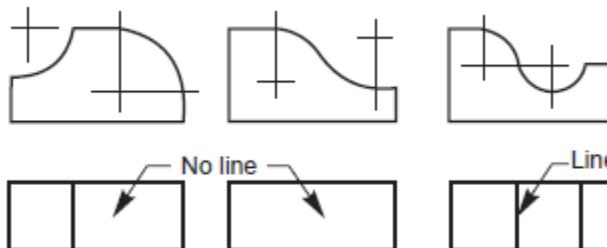


Fig. 5.9 Representation of tangential curved surfaces

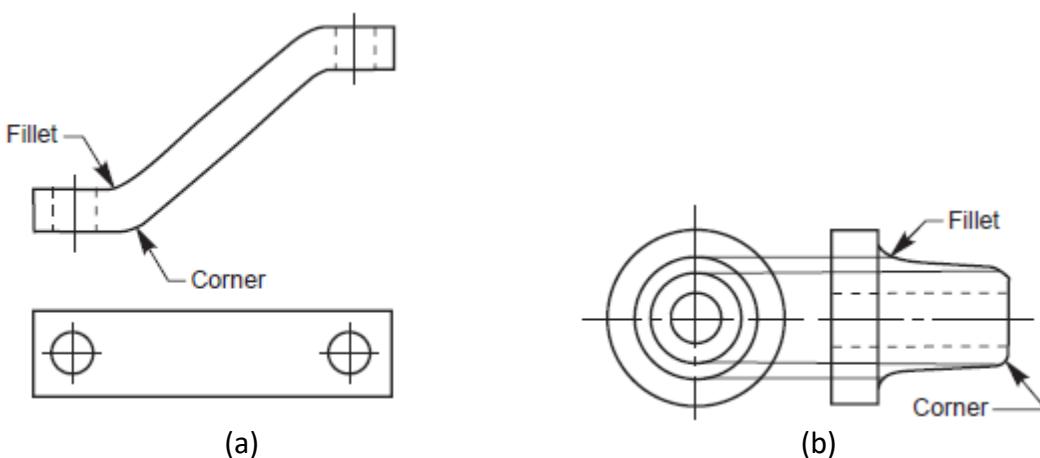


Fig. 5.10 Representation of corners and fillets

If true projection is followed in drawing the view of an object containing fillets and rounds; it will result in misleading impression. In conventional practice, fillets and rounds are represented by lines called run outs. The run outs are terminated at the point of tangency.

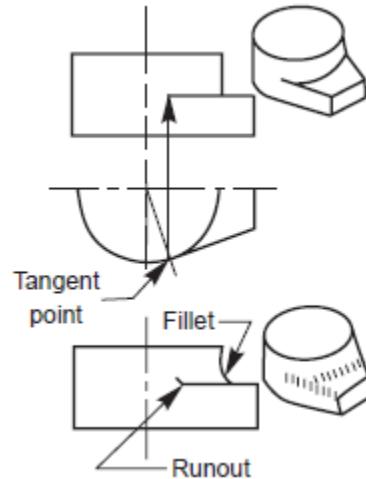


Fig. 5.11 Run outs

Selection of Views

For describing any object completely through its orthographic projections, it is important to select a number of views. The number of views required to describe any object will depend upon the extent of complexity involved in it. The higher the symmetry, the lesser the number of views required.

1. One View Drawings

Some objects with cylindrical, square or hexagonal features or, plates of any size with any number of features in it may be represented by a single view. In such cases, the diameter of the cylinder, the side of the square, the side of the hexagon or the thickness of the plate may be expressed by a note or abbreviation. Square sections are indicated by light crossed diagonal lines. Figure shows some objects which may be described by one-view drawings.

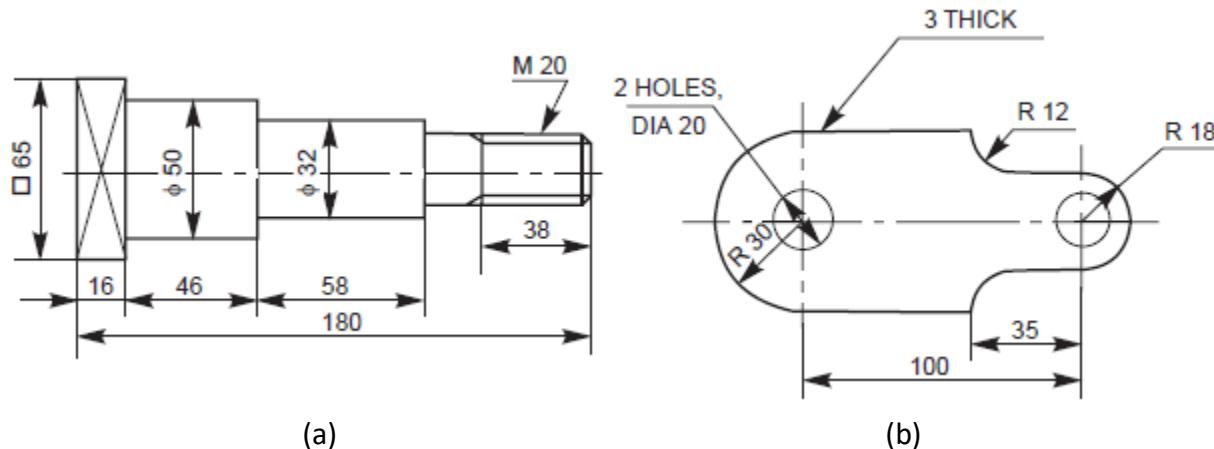


Fig. 5.12 One View Drawings

2. Two – View Drawings

Some objects which are symmetrical about two axes may be represented completely by two views. Normally, the largest face showing most of the details of the object is selected for drawing the view from the front. The shape of the object then determines whether the second view can be a view from above or a side view. Figure shows the example of two-view drawings.

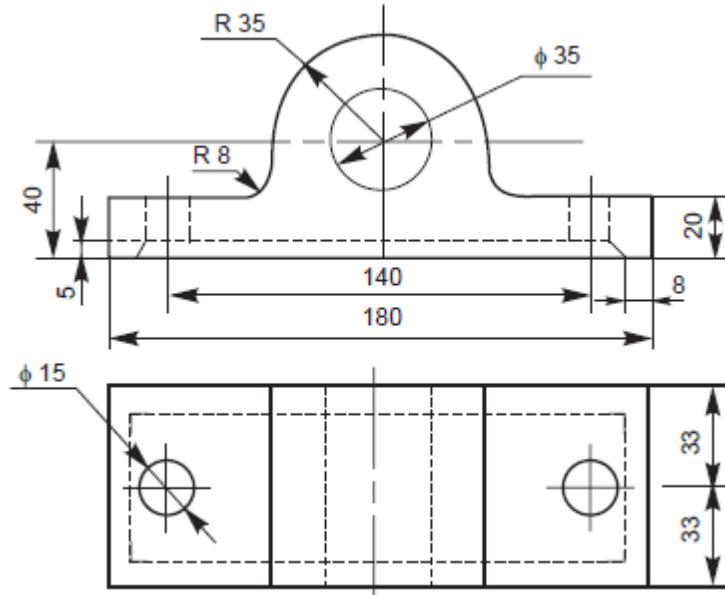


Fig. 5.13 Two – View Drawings

3. Three – View Drawings

In general, most of the objects consisting of either a single component or an assembly of a number of components are described with the help of three views. In such cases, the views normally selected are the views from the front, above and left or right side. Figure shows an object and its three necessary views.

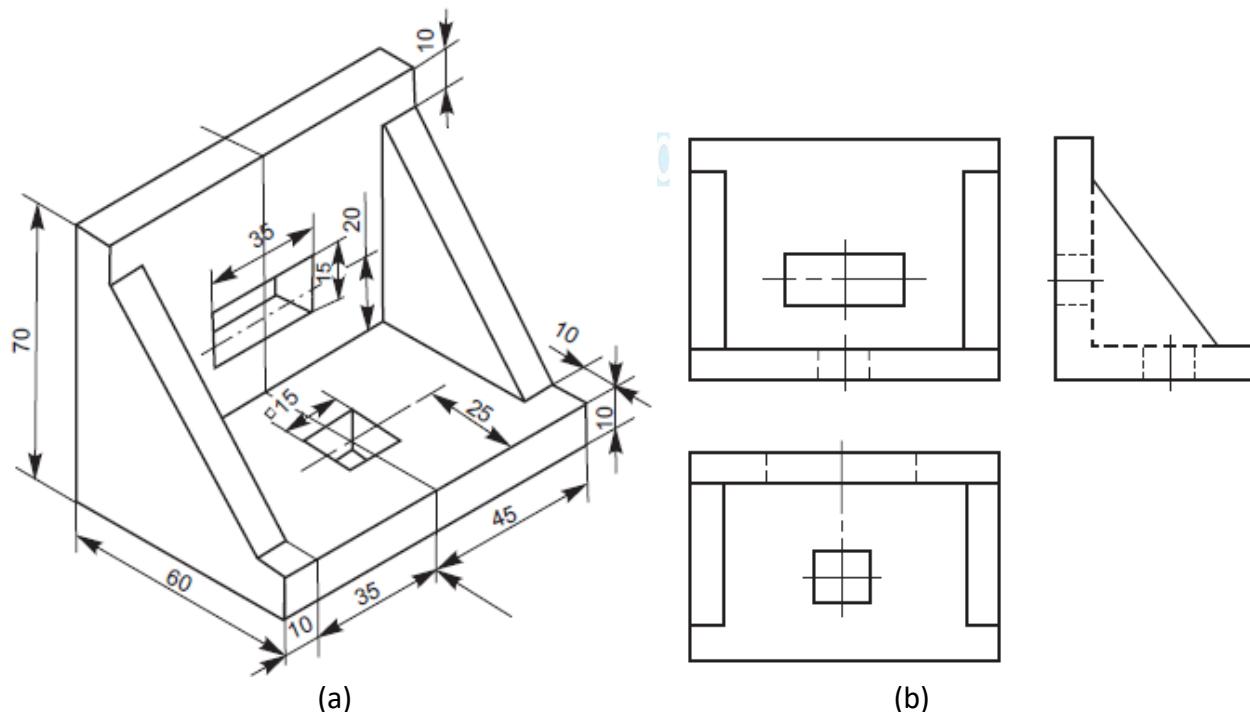


Fig. 5.14 Three – View Drawing

Examples

NOTE: - For all the examples given, the following may be noted: Arrow indicates the direction to obtain the view from the front.

1.

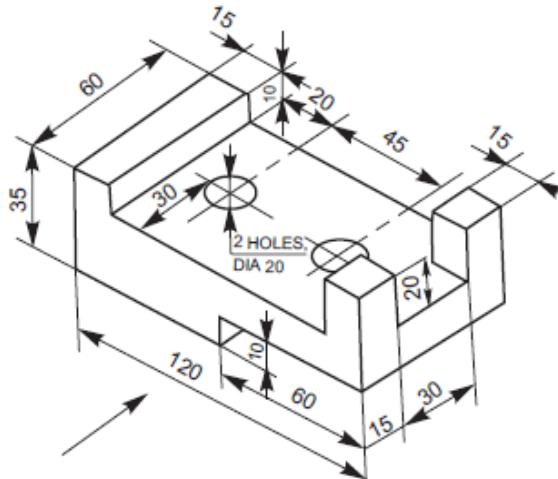


Fig. 5.15

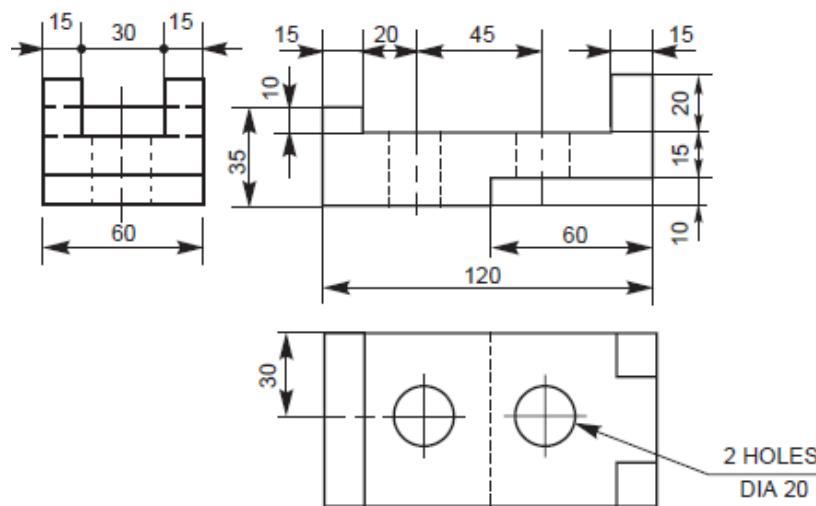
Solution: -

Fig. 5.16

2.

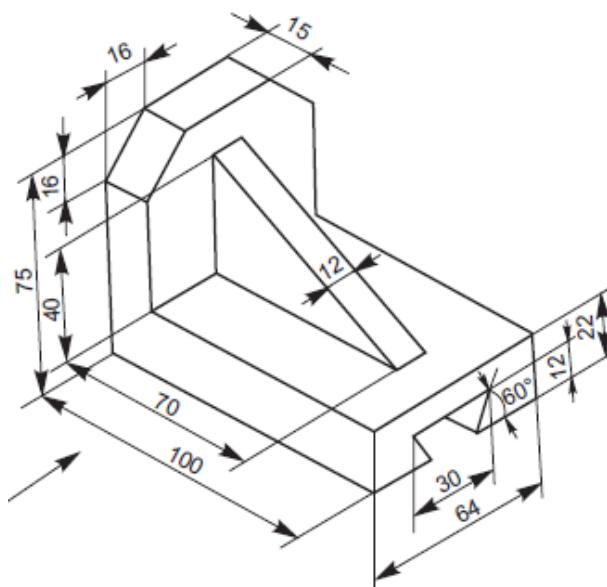


Fig. 5.17

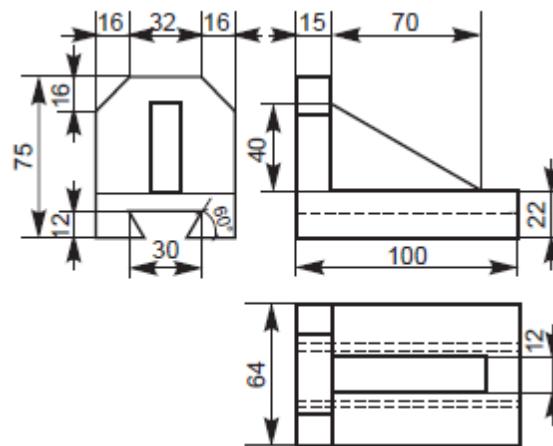
Solution: -

Fig. 5.18

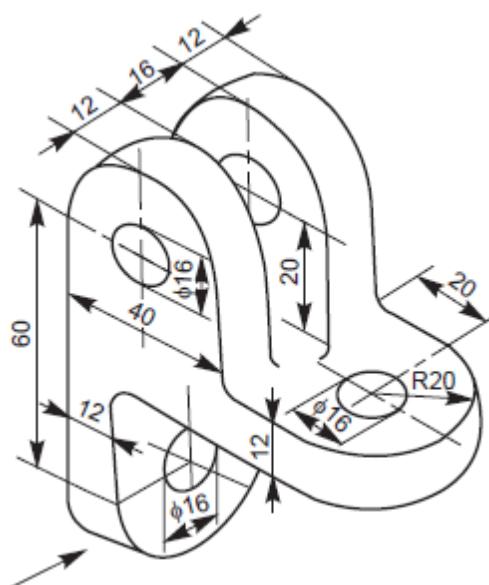
3.

Fig. 5.19

Solution: -

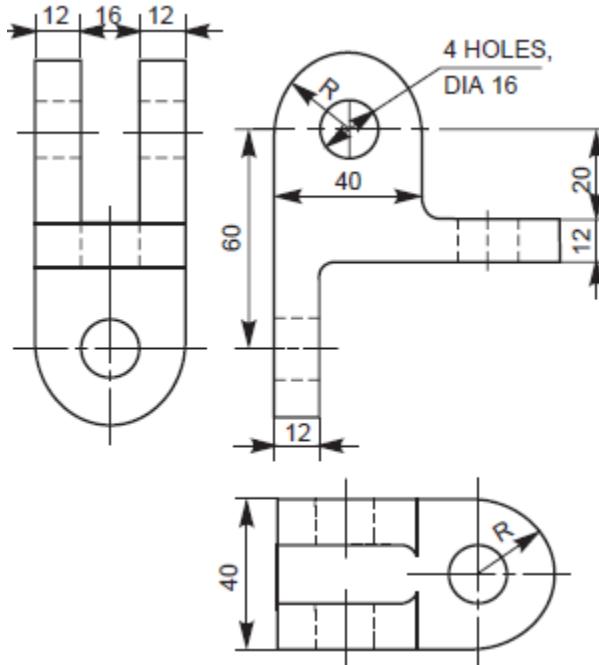


Fig. 5.20

Isometric projections: Isometric scales, isometric views of Simple objects.

Introduction

Pictorial projections are used for presenting ideas which may be easily understood by persons even without technical training and knowledge of multi-view drawing. The Pictorial drawing shows several faces of an object in one view, approximately as it appears to the eye.

Principle of Isometric Projections

It is a pictorial orthographic projection of an object in which a transparent cube containing the object is tilted until one of the solid diagonals of the cube becomes perpendicular to the vertical plane and the three axes are equally inclined to this vertical plane.

Isometric projection of a cube in steps is shown in Fig. Here ABCDEFGH is the isometric projection of the cube.

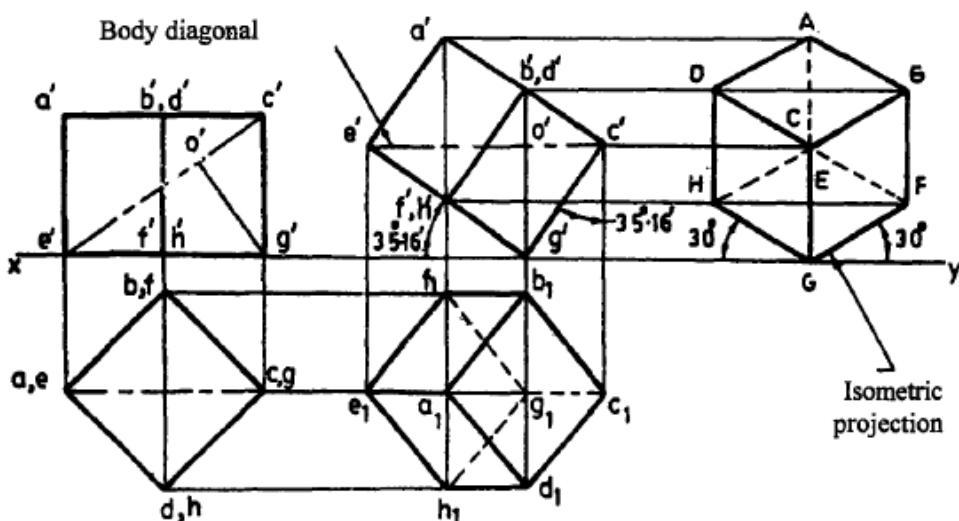


Fig. 5.21 Principle of Isometric Projection

The front view of the cube, resting on one of its corners (G) is the isometric projection of the cube.

Isometric Scale

In the isometric projection of a cube shown in Fig., the top face ABCD is sloping away from the observer and hence the edges of the top face will appear fore-shortened. The true shape of the triangle DAB is represented by the triangle DPB.

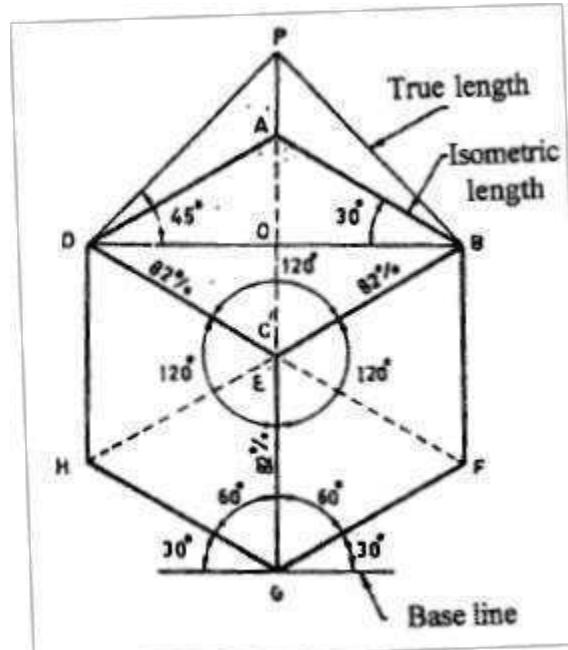


Fig. 5.22 An isometric Cube

The extent of reduction of an isometric line can be easily found by construction of a diagram called isometric scale. For this, reproduce the triangle DPA as shown in Fig. Mark the divisions of true length on DP. Through these divisions draw vertical lines to get the corresponding points on DA. The divisions of the line DA give dimensions to isometric scale.

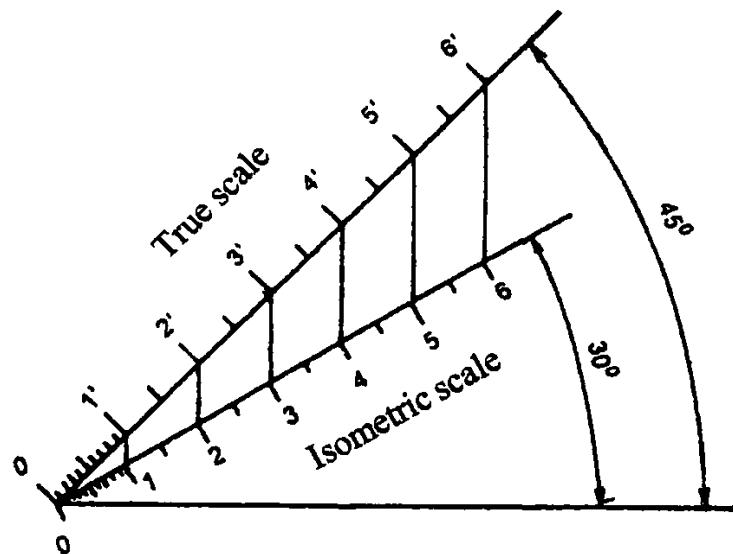


Fig. 5.23 Isometric Scale

From the triangle ADO and PDO in Fig., the ratio of the isometric length to the true length, i.e., $DA/DP = \cos 45^\circ/\cos 30^\circ = 0.816$

The isometric axes are reduced in the ratio 1:0.816 i.e. 82% approximately.

Lines in Isometric Projection

The following are the relations between the lines in isometric projection which are evident from Fig..

1. The lines that are parallel on the object are parallel in the isometric projection.
2. Vertical lines on the object appear vertical in the isometric projection.
3. Horizontal lines on the object are drawn at an angle of 30° with the horizontal in the isometric projection.
4. A line parallel to an isometric axis is called an isometric line and it is fore shortened to 82%.
5. A line which is not parallel to any isometric axis is called non-isometric line and the extents of foreshortening of non-isometric lines are different if their inclinations with the vertical planes are different.

Isometric Projection

Figure (a) shows a rectangular block in pictorial form and Fig. (b), the steps for drawing an isometric projection using the isometric scale.

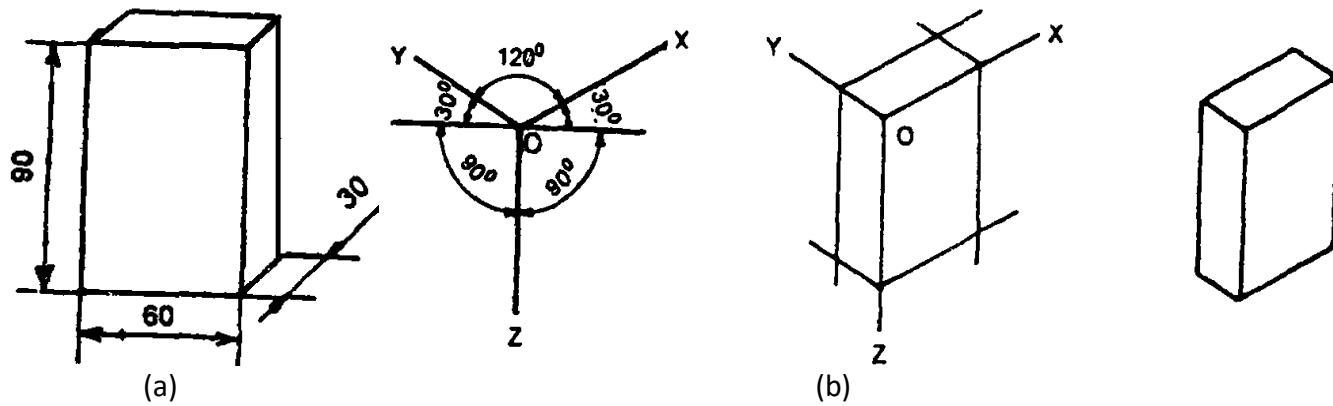


Fig. 5.24 Developing Isometric Projection

Isometric Drawing

Drawings of objects are seldom drawn in true isometric projections, as the use of an isometric scale is inconvenient. Instead, a convenient method in which the foreshortening of lengths is ignored and actual or true lengths are used to obtain the projections, called isometric drawing or isometric view is normally used. This is advantageous because the measurement may be made directly from a drawing.

The isometric drawing of figure is slightly larger (approximately 22%) than the isometric projection. As the proportions are the same, the increased size does not affect the pictorial value of the representation and at the same time, it may be done quickly. Figure shows the difference between the isometric drawing and isometric projection.

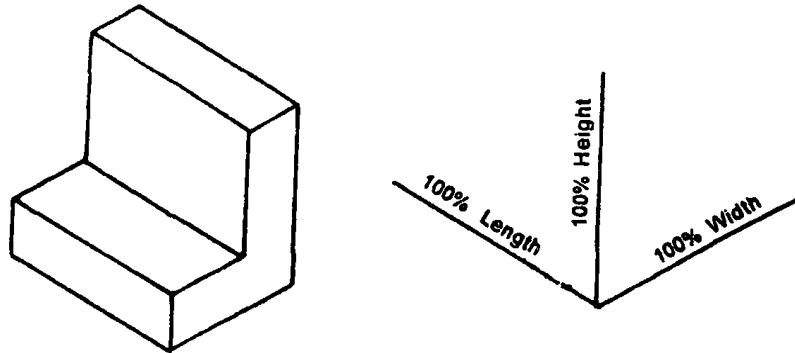


Fig. 5.25 (a) Isometric Drawing

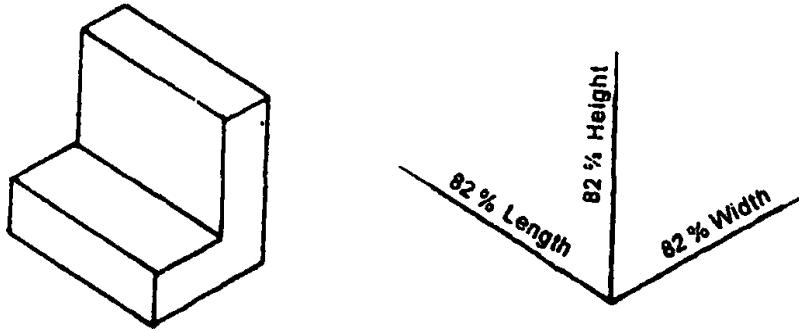


Fig. 5.26 (b) Isometric Projection

Steps to be followed to make isometric drawing from orthographic views are given below (Fig.).

1. Study the given views and note the principal dimensions and other features of the object.
2. Draw the isometric axes (a).
3. Mark the principal dimensions to-their true values along the isometric axes (b).
4. Complete the housing block by drawing lines parallel to the isometric axes and passing through the above markings (e).
5. Locate the principal corners of all the features of the object on the three faces of the housing block (d).
6. Draw lines parallel to the axes and passing through the above points and obtain the isometric drawing of the object by darkening the visible edges (e).

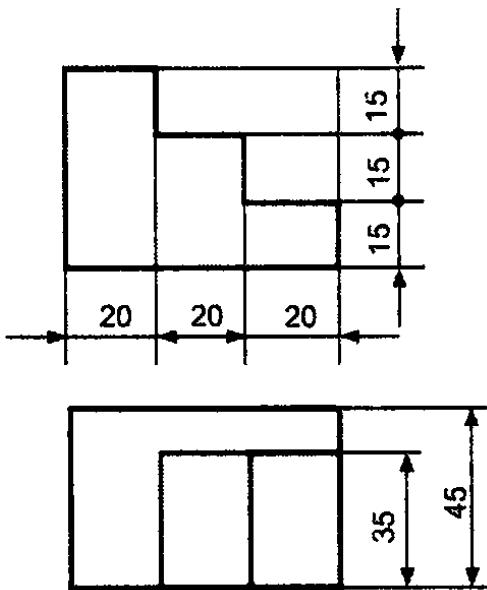
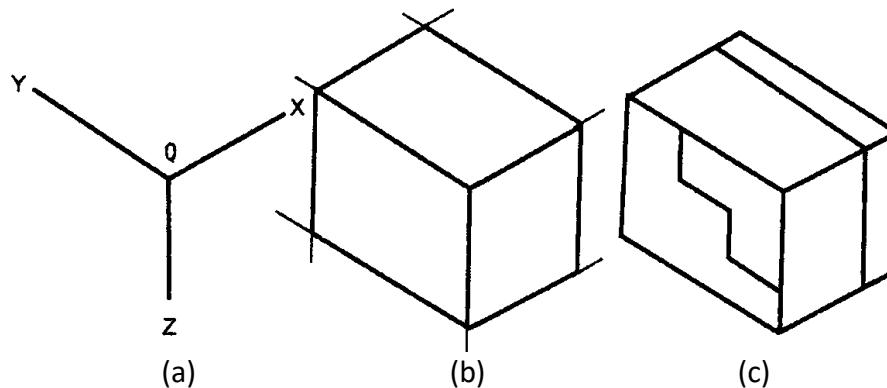
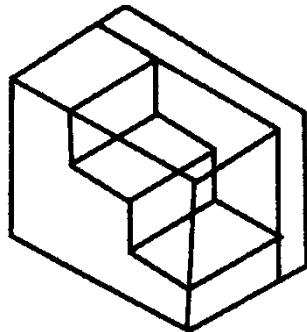


Fig. 5.27 (A) Orthographic View





(d)

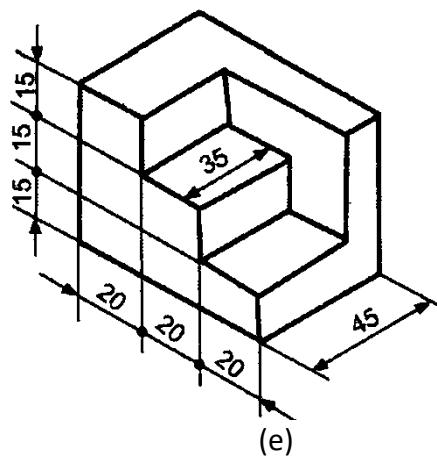


Fig. 5.28 (B) Isometric View

Non-Isometric Lines

In an isometric projection or drawing, the lines that are not parallel to the isometric axes are called non-isometric lines. These lines obviously do not appear in their true length on the drawing and cannot be measured directly. These lines are drawn in an isometric projection or drawing by locating their end points. Figure shows the isometric drawing of an object containing non isometric lines from the given orthographic views.

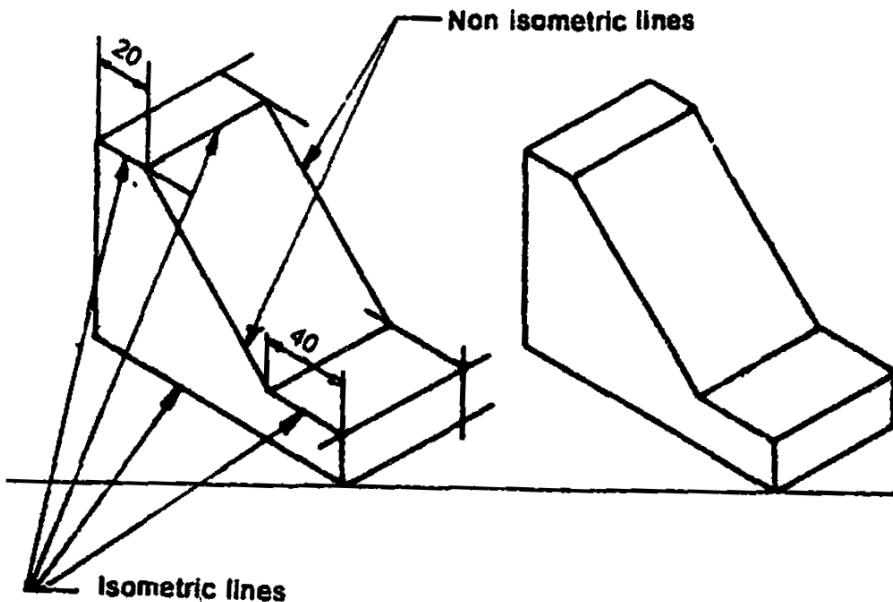


Fig. 5.29

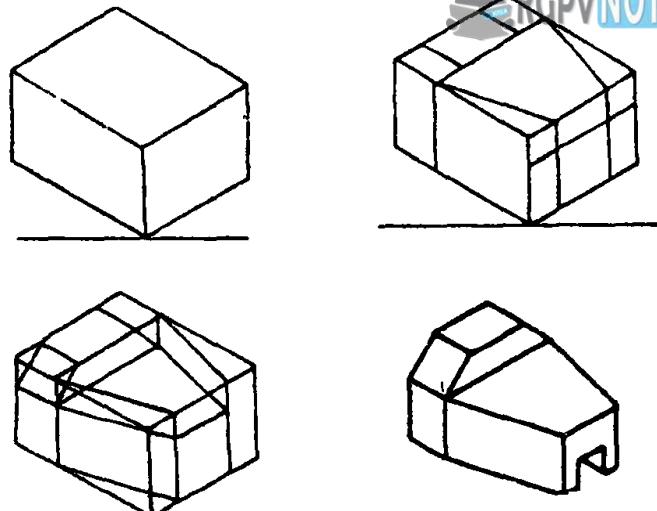
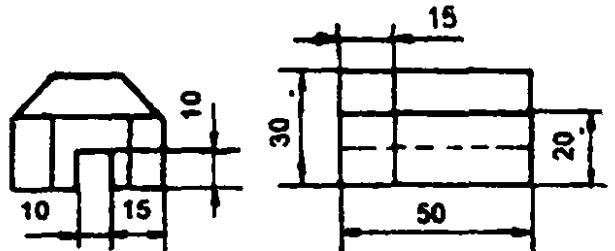
Methods of Constructing Isometric Drawing

The methods used are:

1. Box method.
 2. Off-set method.

Box Method

When an object contains a number of non-isometric lines, the isometric drawing may be conveniently constructed by using the box method. In this method, the object is imagined to be enclosed in a rectangular box and both isometric and non-isometric lines are located by their respective points of contact with the surfaces and edges of the box.



(a)

(b)

Fig. 5.30

Off-set Method

Off-set method of making an isometric drawing is preferred when the object contains irregular curved surfaces. In the off-set method, the curved feature may be obtained by plotting the points on the curve, located by the measurements along isometric lines. Figure illustrates the application of this method.

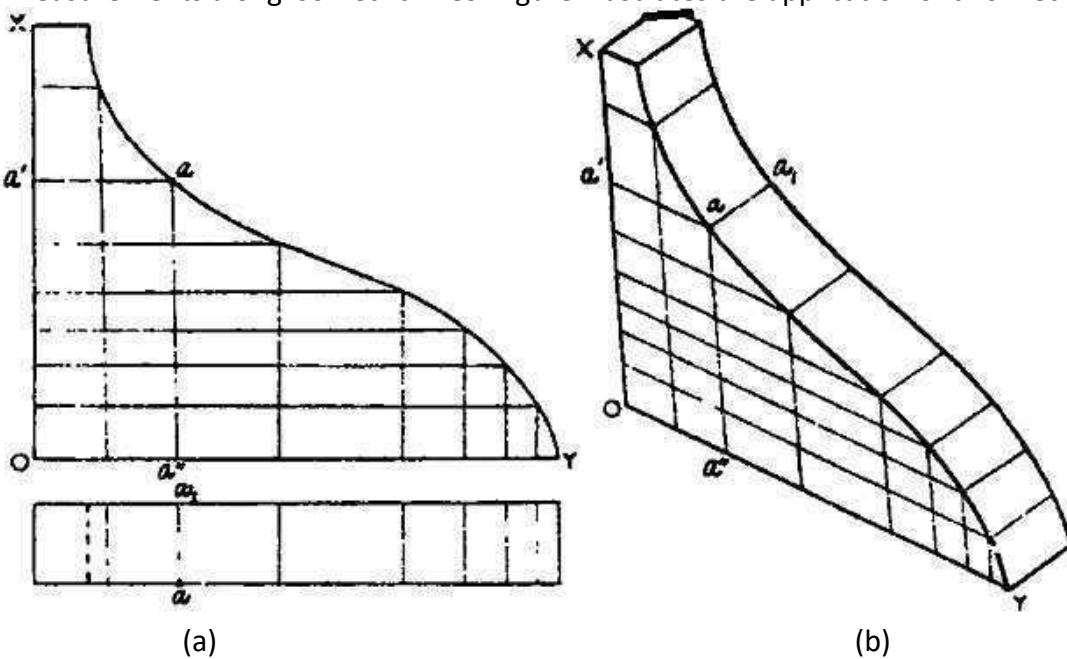


Fig. 5.31

Isometric Projection of Planes

Problem: Draw the isometric projection of a rectangle of 100mm and 70mm sides if its plane is (a) Vertical and (b) Horizontal.

Construction:-

1. Draw the given rectangle ABCD as shown in Fig. (a).

Note:

- (i) In the isometric projection, vertical lines are drawn vertical and the horizontal lines are drawn inclined 30° to the base line.
 - (ii) As the sides of the rectangle are parallel to the isometric axes they are fore-shortened to approximately 82% in the isometric projections.
- Hence $AB = CD = 1000 \times 0.82\text{mm} = 82\text{mm}$. Similarly, $BC = AD = 57.4\text{ mm}$.

(a) When the plane is vertical:

2. Draw the side AD inclined at 30° to the base line as shown in Fig. b and mark $AD = 57.4\text{mm}$.

3. Draw the verticals at A and D and mark off $AB = DC = 82\text{mm}$ on these verticals.

4. Join BC which is parallel to AD.

ABCD is the required isometric projection. This can also be drawn as shown in Fig. c. Arrows show the direction of viewing.

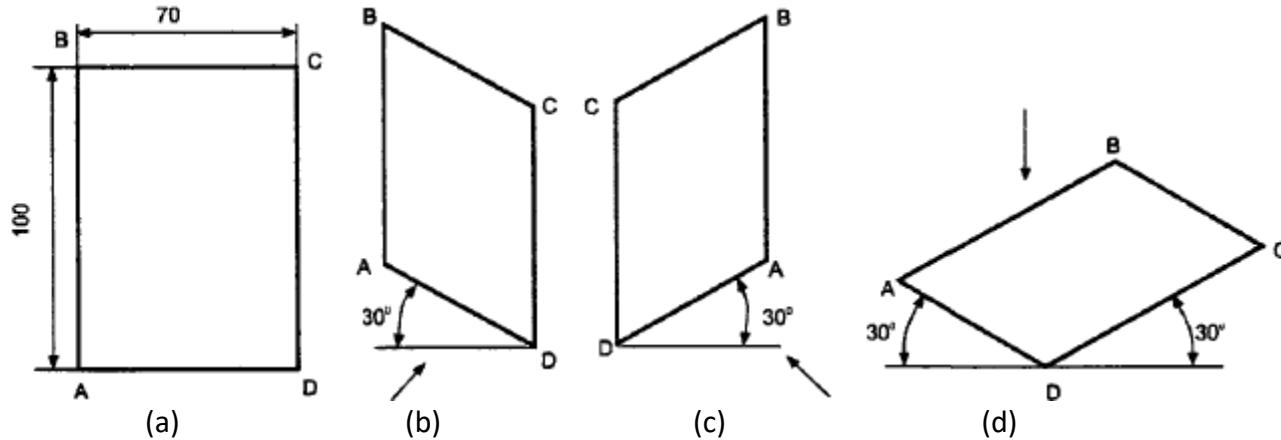


Fig. 5.32

(b) When the plane is horizontal.

5. Draw the sides AD and DC inclined at 30° to be base line and complete the isometric projection ABCD as shown in Fig. d. Arrow at the top shows the direction of viewing.

Problem: Figure shows the projection of a pentagonal plane. Draw the isometric drawing of the plane (i) when the surface is parallel to V.P. and (ii) parallel to H.P.

Construction: -

1. Enclose the given pentagon in a rectangle 1234.

2. Make the isometric drawing of the rectangle 1234 by using true lengths.

3. Locate the points A and B such that $l_a = l_A$ and $l_b = l_B$.

4. Similarly locate point C, D and E such that $2c = 2C$, $3d = 3D$ and $e4 = E4$.

5. ABCDE is the isometric drawing of the pentagon.

6. Following the above principle of construction fig. c can be

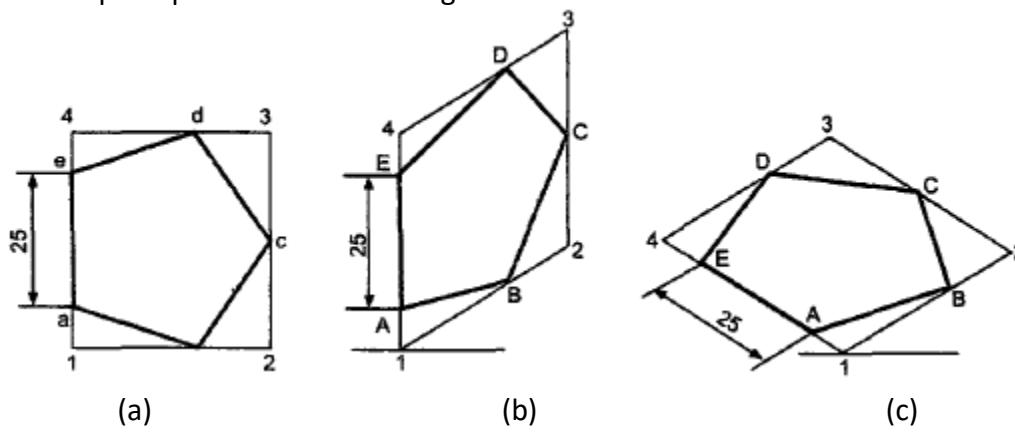
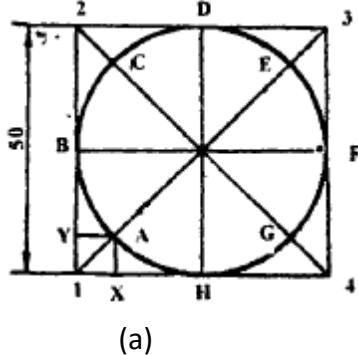


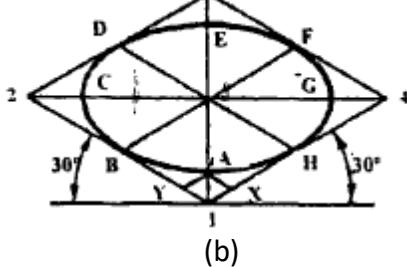
Fig. 5.33

Problem: Draw the isometric view of a circular plane of diameter 60mm whose surface is (a) Horizontal, (b) Vertical.

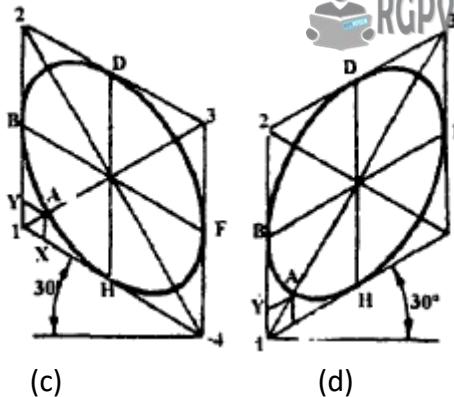
Construction: - Using the method of points



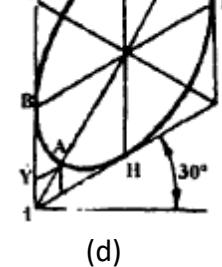
(a)



(b)



(c)



(d)

Fig. 5.34

1. Enclose the circle in a square 1-2-3-4 and draw diagonals, as shown in Fig. 5.34 a. Also draw lines YA horizontally and XA vertically.

To draw the isometric view of the square 1-2-3-4 as shown in Fig. 5.34 b.

2. Mark the mid points of the sides of the square as B, D, F and H.

3. Locate the points X and Y on lines 1-4 and 1-2 respectively.

4. Through the point X, draw AX parallel to line 1-2 to get point A on the diagonal 1-3. The point A can be obtained also by drawing YA through the point Y and parallel to the line 1-4. Similarly obtain other points C, E and G

6. Draw a smooth curve passing through all the points to obtain the required isometric view of the horizontal circular plane.

7. Similarly obtain isometric view of the vertical circular plane as shown in Fig. 5.34 c and d.

Problem: Draw the isometric projection of a circular plane of diameter 60mm whose surface is (a) Horizontal and (b) Vertical-use Four-centre method



Construction: - Using Four – Centre Method

1. Draw the isometric projection of the square 1-2-3-4 (rhombus) whose length of side is equal to the isometric length of the diameter of the circle = 0.82×60 .

2. Mark the mid points A, B, C and D of the four sides of the rhombus. Join the points 3 and A. This line intersects the line 2-4 joining the point 2 and 4 at M. Similarly obtain the intersecting point N.

3. With centre M and radius = MA draw an arc A B. Also draw an arc C D with centre N.

4. With centre 1 and radius = 1C, draw an arc B C. Also draw the arc A D.

5. The ellipse ABCD is the required isometric projection of the horizontal circular plane (Fig. 5.35 a).

6. Similarly obtain the isometric projection in the vertical plane as shown in Fig. 5.35 b & c.

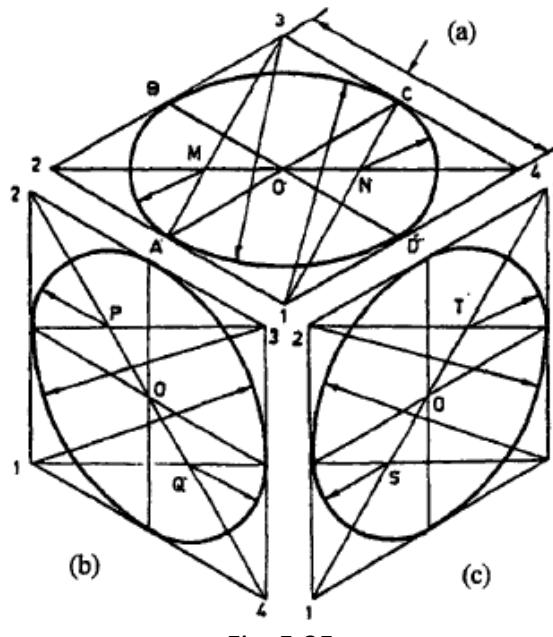


Fig. 5.35

Isometric Projection of Prisms

Problem: Draw the isometric view of a pentagonal prism of base 60 mm side, axis 100 mm long and resting on its base with a vertical face perpendicular to V.P.

Construction: -

1. The front and top views of the prism are shown in Fig. a.
2. Enclose the prism in a rectangular box and draw the isometric view as shown in Fig. b using the box method.

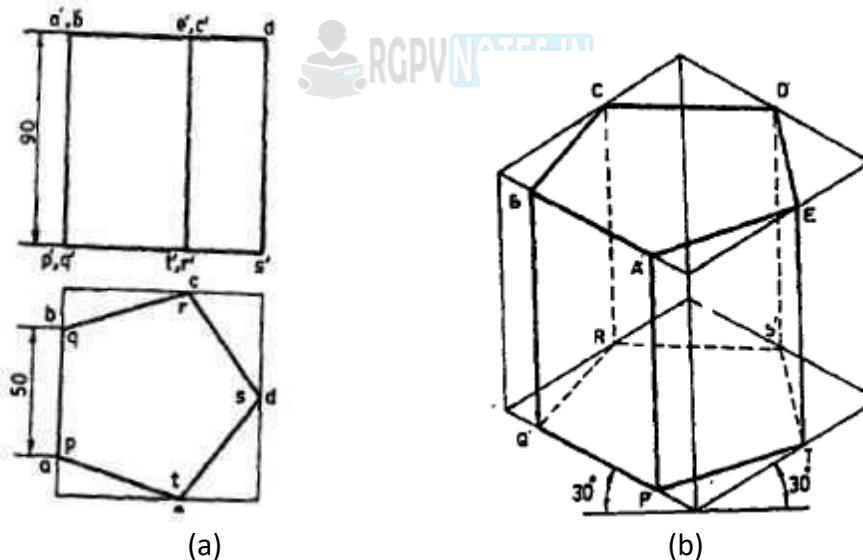
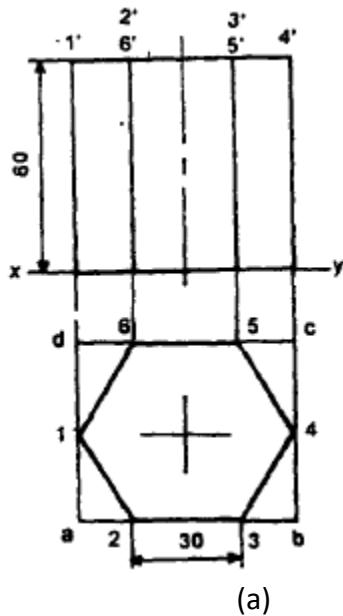


Fig. 5.36 Isometric Drawing of a Pentagonal Prism

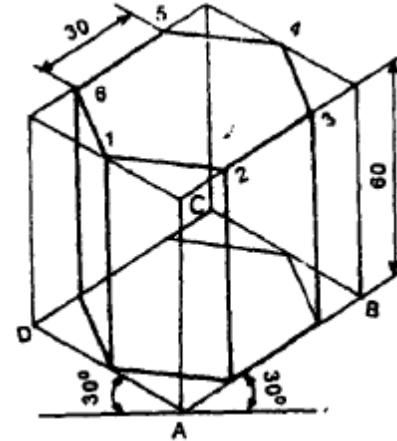
Problem: A hexagonal prism of base of side 30 mm and height 60 mm is resting on its base on H.P. Draw the isometric drawing of the prism.

Construction: -

1. Draw the orthographic views of the prism as shown in Fig. a.
2. Enclose the views in a rectangle (i.e. the top view –base and front views).
3. Determine the distances (off-sets) of the corners of the base from the edges of the box.
4. Join the points and darken the visible edges to get the isometric view.



(a)



(b)

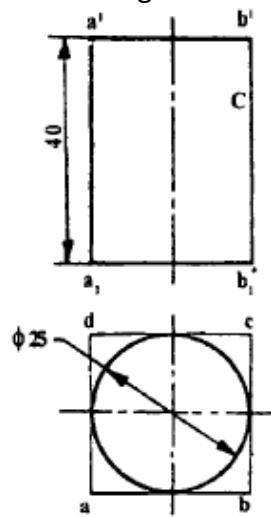
Fig. 5.37 Isometric Drawing of a Hexagonal Prism

Isometric Projection of Cylinder

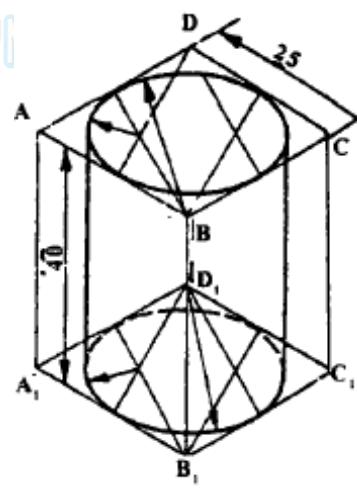
Problem: Make the isometric drawing of a cylinder of base diameter 20 mm and axis 35 mm long.

Construction: -

1. Enclose the cylinder in a box and draw its isometric drawing.
2. Draw ellipses corresponding to the bottom and top bases by four centre method.
3. Join the bases by two common tangents.



(a)



(b)

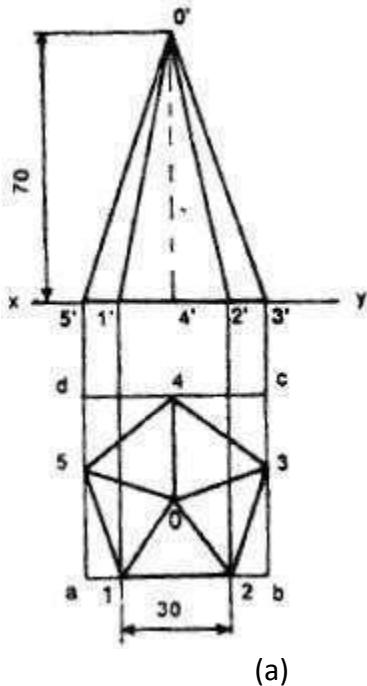
Fig. 5.38 Isometric Drawing of Cylinder

Isometric Projection of Pyramid

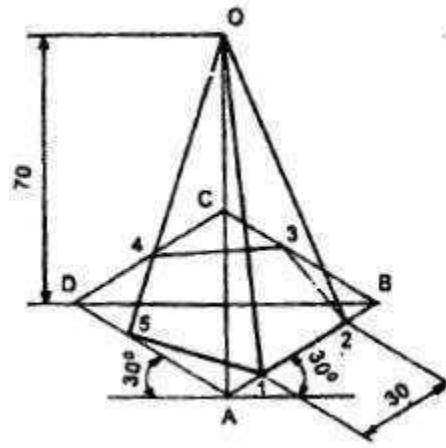
Problem: A pentagonal pyramid of side of base 30 mm and height 70 mm is resting with its base on H.P. Draw the isometric drawing of the pyramid.

Construction: -

1. Draw the projections of the pyramid (Fig. a).
2. Enclose the top view in a rectangle abcde and measure the off-sets of all the corners of the base and the vertex.
3. Draw the isometric view of the rectangle ABCD.
4. Using the off-sets locate the corners of the base 1, 2, etc. and the vertex o.
5. Join 0-1, 0-2, 0-3, etc. and darken the visible edges and obtain the required view.



(a)



(b)

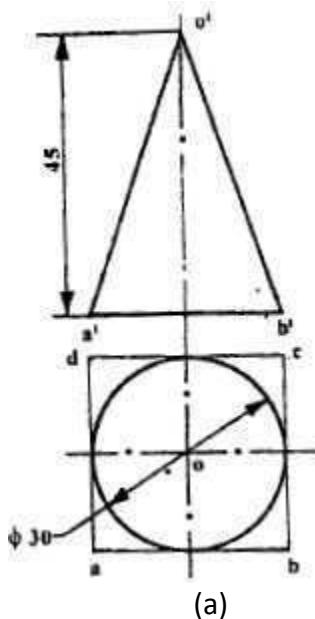
Fig. 5.39 Isometric Drawing of Pentagonal Pyramid

Isometric Projection of Cone

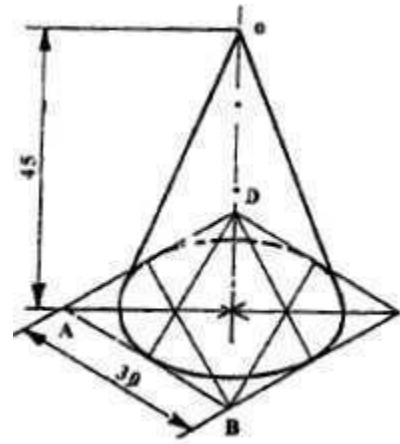
Problem: Draw the isometric drawing of a cone of base diameter 30 mm and axis 50 mm long.

Construction: - Using offset method

1. Enclose the base of the cone in a square (Fig. a).
2. Draw the ellipse corresponding to the circular base of the cone.
3. From the centre of the ellipse draw a vertical centre line and locate the apex at a height of 50 mm.
4. Draw the two outer most generators from the apex to the ellipse and complete the drawing.



(a)



(b)

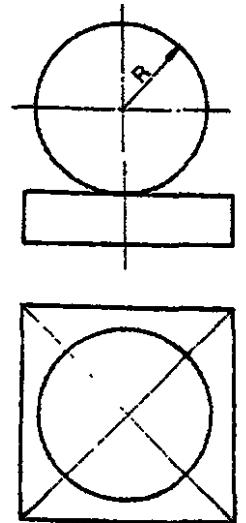
Fig. 5.40

Examples: -

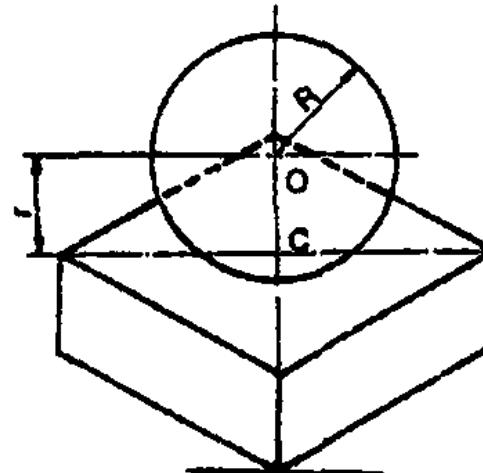
The orthographic projections and the isometric projections of some solids and machine components: -

Note here (a) will be question and (b) will be answer.

1.



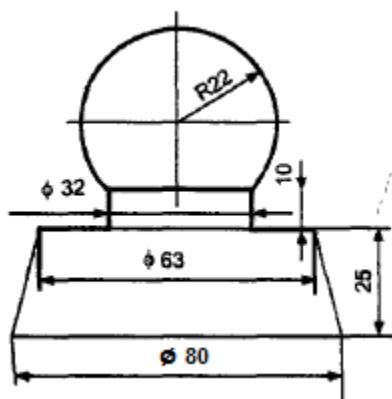
(a)



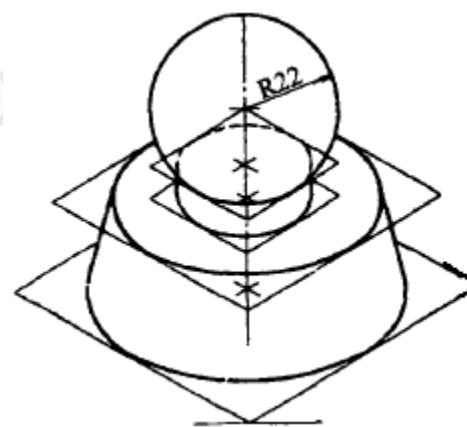
(b)

Fig. 5.41

2.

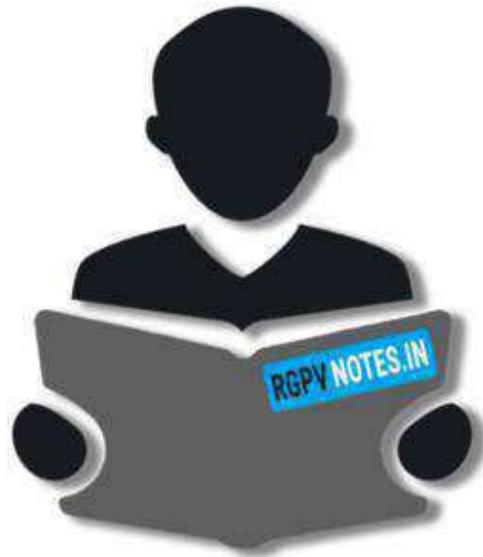


(a)



(b)

Fig. 5.42



RGPVNOTES.IN

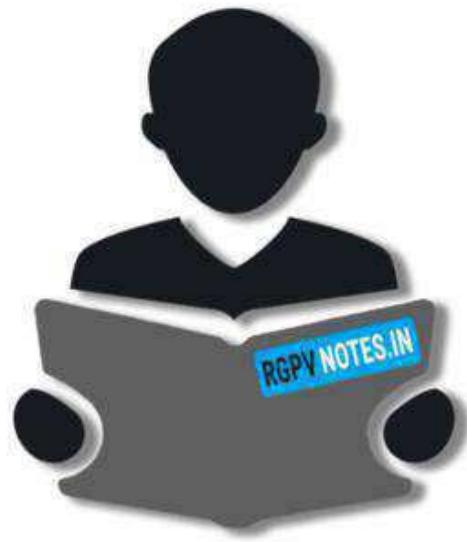
We hope you find these notes useful.

You can get previous year question papers at
<https://qp.rgpvnotes.in> .

If you have any queries or you want to submit your
study notes please write us at
rgpvnotes.in@gmail.com



LIKE & FOLLOW US ON FACEBOOK
facebook.com/rgpvnotes.in



RGPVNOTES.IN

Program : **B.Tech**

Subject Name: **Engineering Chemistry**

Subject Code: **BT-101**

Semester: **1st**



LIKE & FOLLOW US ON FACEBOOK
facebook.com/rgpvnotes.in

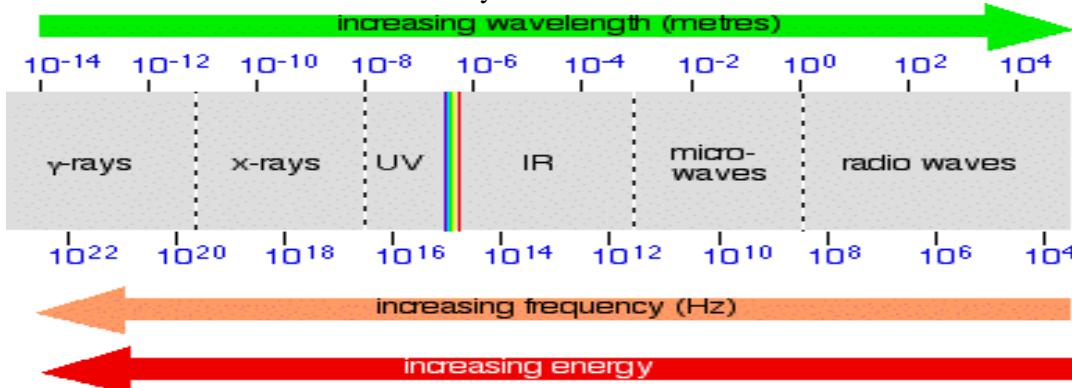
UNIT V: INSTRUMENTAL TECHNIQUES IN CHEMICAL ANALYSIS

UV SPECTROSCOPY

INTRODUCTION

The spectroscopic methods have proved to be extensively useful for qualitative analysis identification and structural elucidation of organic compounds. Spectroscopic methods are advantageous over other methods because

- (i) They are quick
- (ii) They are sensitive
- (iii) Information obtained by them is in the form of permanent record
- (iv) The informations obtained by these methods are highly reliable
- (v) Very small quantity of substance is required for the determination of spectra and even the quantity used can also be recovered after the analysis.



Origin of Electronic Spectra – Spectroscopy is the term applied for the branch of science which deals with the study of resolution of visible radiations into its component wave lengths. Now, the term is broadened and is applied to studies involving entire electromagnetic spectrum. Electromagnetic radiations consists of waves of energy and their parameters as:

- (i) **Wavelength** – (λ) It is the distance between two consecutive peaks or crests.
- (ii) **Wave Number** – (V) It is the number of waves per cm.
- (iii) **Frequency** – (V) It is the number of waves per second.

Different types of electronic spectra are – Emission Spectra, Absorption Spectra.

Emission spectra- when a substance is subjected to intense heat or to an electric discharge, its atoms & molecules absorb energy & get excited. These excited species, on returning to the ground state may emit radiation which on passing through a prism gives rise to a spectrum of atoms appears as bright lines on a dark background where as spectra of molecules appears as band, when white light is passed through a prism, we get a continuous spectrum of seven colours and sodium light gives a line spectrum [two-D-line] (yellow) of define wave length both are emission spectra.

Absorption spectra- when white light is passed through yellow sodium flame before reaching the prism, we get a continuous spectrum with two black lines in place of yellow lines obtained from sodium light. Here sodium flame has absorbed two wavelengths from white light which it itself emits. This is absorption spectrum.

During the absorption, some molecules falling in the path of incident beam colloid with photons of energy exactly equal to the difference in energy between the ground and excited states of the molecules. Thus a spectroscopic technique, qualitative or quantitative if it depends upon the measurement of an absorption spectrum is called absorption spectroscopy or molecular spectroscopy.

Principle- consider a molecule of a compound [X] have only two energy levels E_1 and E_2 the energy difference ϵE is determined as $\epsilon E = E_2 - E_1 = h\nu$ or hc/λ

Energy absorbed by each molecule is given by $\epsilon E = N hc/\lambda$

Where

$$\begin{aligned}
 h &= \text{Plank's constant} = 6.63 \times 10^{-34} \\
 c &= \text{Velocity of electromagnetic radiations} = 3 \times 10^8 \text{ ms}^{-1} \\
 N &= 6.02 \times 10^{23} \text{ mol}^{-1}
 \end{aligned}$$

UNIT V: INSTRUMENTAL TECHNIQUES IN CHEMICAL ANALYSIS

$$\lambda = \text{wavelength}$$

The light of wavelength λ has photon of energy just equal to the energy difference between the two energy levels. In actual practice, molecules of compound have large number of energy levels and very small but definite amounts of energy is needed for transition between some of these while extremely large amounts of energy are involved between some other transitions. The following transitions are brought about by absorption of radiant energy by a molecule –

- (i) **Electronic Transition** – Electrons are jumped to higher energy levels by the absorption of energy.
- (ii) **Vibrational Transition** – Stretching, contracting and bending of covalent bond may occur due to absorption of energy.
- (iii) **Rotational Transition** – Due to the absorption of radiant energy change in the rotational energy of the molecule takes place.

Bond	A	B
Bond Stretching	A	B
Contracting	A	B
Bending	A	B

Any wavelength of radiations absorbed by a molecule is determined by the changes in the electronic, Vibrational or rotational energy levels, permissible for it and its atoms.

- (i) High-energy radiations as UV or visible are required to bring about electronic transition.
- (ii) The low energy radiations as IR are required for Vibrational transitions.
- (iii) The radiation of far infrared regions brings about rotational transitions.

ABSORBANCE

The intensity of absorption is related to the number of photons absorbed by the molecules usually some photons are absorbed by the molecules. The fraction of photon absorbed by the molecules at given frequency depends upon

- (i) The nature of absorbing molecule.
- (ii) The concentration of molecules. The higher the concentration the more molecules are present to absorb the photons.
- (iii) The length of the path of the radiation through the material. The longer path, the larger number of molecules exposed and hence greater is the probability that a given photon will be absorbed.

Laws of Absorbance

- (i) **Lamberts Law** – If a monochromatic light passes through a transparent medium, the rate of decrease in intensity with the thickness of medium is proportional to the intensity of the incident light i.e. the intensity of the emitted light decreases exponentially as the thickness of the absorbing medium increases arithmetically.

$$dI/I = -k dx \quad dI/I = -k dx$$

dI = Change in transmitted light

I = Intensity of light

dx = Small thickness of medium

k = Proportionality constant

On integrating his equation between $I = I_0$ limits at $x = 0$ and $I = I$ at $x = L$ we get

- (ii) **Beers Law** – The intensity of beam of monochromatic light decreases exponentially as the concentration of the absorbing substance increases arithmetically.

$$I_t = I_0 e^{-k'c}$$

According to Beers law, the absorbance at any particular wave length is directly proportional to the number of absorbing molecules. If the solution contains more than one type of absorbing species, the total absorbance will be the sum of the absorbance of all the species provided they do not interact chemically.

$$\begin{aligned}
 I_t &= I_0 e^{-k'c} \\
 &= I_0 e^{-0.47343 k'c} \\
 &= I_0 e^{-k'c} 3/4 \text{ (Beer's Law)}
 \end{aligned}$$

UNIT V: INSTRUMENTAL TECHNIQUES IN CHEMICAL ANALYSIS

- It = Intensity of Transmitted light
 Io = Intensity of incident light
 k = Proportionality constant
 C = Concentration (Molar) of medium

(iii) Beer's Lambert's Law – Electronic spectra of organic compound are most frequently recorded in solution when a monochromatic light of intensity Io is passed through a solution of concentration C Molar, then intensity of Transmitted light It changes. The probability of absorption of r addition is determined by a mathematical expression given by Beer's lamberts law which expresses the relationship between amount of light absorbed and -

- (i) Concentration of the solution
- (ii) Length of the solution through which light passes

$$\begin{aligned} \text{In } \frac{I}{I_0} &= -K L \quad \text{or } I_t = I_0 e^{-KL} \quad \text{--- (i) Lambert's Law} \\ I_t &= I_0 e^{-K \cdot C} \end{aligned}$$

From lambert's Law

$$\begin{aligned} \text{In } \frac{I}{I_0} &= -K l \\ 2.303 \log \frac{I}{I_0} &= -K l \end{aligned}$$

$$\log \frac{I_0}{I} = K / 2.303 c l = \epsilon c l = A$$

$\epsilon = K/2.303$ and is called molar absorptivity coefficient

From Beer's law

$$\begin{aligned} \text{Log } \frac{I_0}{I} &= A \text{ Absorbance} \\ I_t &= I_0 e^{-K \cdot C} = I_0 10^{-K \cdot C} \end{aligned}$$

By combining equation (i) and (ii) i.e. lambert's law and Beer's law.

$$I_t = I_0 . 10^{-\epsilon c l}$$

$$\text{Where } \epsilon = K/2.303$$

$$\log \frac{I_0}{I} = \epsilon c l$$

$$\text{Where } I_t/I_0 = A$$

$$A = \epsilon x c x l \text{ or } \epsilon = A/c l$$

I_0 = Intensity of incident light

I_t or I = Intensity of transmitted light

C = Concentration of absorbing compound in moles per litre

L = Length of sample in cm

ϵ = (Epsilon) is called molar absorptivity co-efficient

A = Absorbance of the solution

$A = \log I_0/I$ or $T = \log I_0/I$ or $T = I/A$

Where T is transmittance of the solution. The wavelength at which a molecule has highest absorption co-efficient [ϵ_{max}] is designated as [λ_{max}]. a spectrum may have several different maxima each with characteristic value of [λ_{max}].

INSTRUMENT USED FOR MEASUREMENT OF ABSORPTION

All chemicals interact with the electromagnetic radiations and due to this there occurs decrease in the intensity of radiant beam. This decrease in intensity is measured by absorption spectroscopic methods.

The various instruments used for measurement of absorption are:

- (i) Colorimeter
- (ii) Absorptionmeter
- (iii) Spectro photometer

UNIT V: INSTRUMENTAL TECHNIQUES IN CHEMICAL ANALYSIS

- (i) **Colorimeter** – It is concerned with the determination of concentration of substance by measurement of relative absorption of light with respect to known concentration of the substance [Absorption in visible region is employed].
- (ii) **Absorptionmeter** – Includes instrument which are useful in other spectral region as well. This measures the ratio of some function of two, of radiant power or two electromagnetic beams.
- (iii) **Spectrometer** – It is an absorption meter used for much narrower bands of wave lengths as produced by monochromator. They can be used in UV, Visible and IR region.

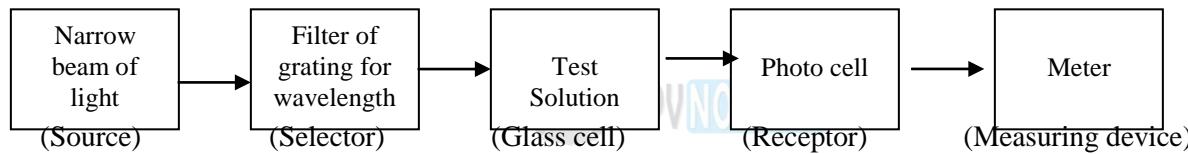
All these instruments have some common features like –

COLORIMETRIC ANALYSIS

The variation of color of the system with change in concentration is the basis of colorimetry. In colorimetry the concentration of a substance is determined by measurement of relative absorption of light with respect to a known concentration of a substance. In visual colorimetry, natural or artificial white light is generally used as a source of light and determinations are made by colorimeter because the method is convenient for the systems in which substances or their solutions are coloured. The eye is replaced by photoelectric cell and intensity of colour is easily measured by photoelectric colorimeter. If the substance is colourless then a suitable complexing agent is used to get a coloured complex (which absorb light in the visible region) for example for the estimation of cuprous ion, complexing agent ammonium hydroxide, is added to get blue coloured solution.

In a colorimeter, a narrow beam of light (of proper wavelength) passes through the solution under test towards a sensitive photocell. Usually colorimeters are provided with arrangements of filters or diffraction grating. Consequently it is possible to select the most appropriate wavelength by choosing a filter grating.

Signal Indicator



The current generated in the photocell is proportional to the amount of light transmitted by the solution. This depends upon depth of colour of substance under test. Thus current from photocell will be greatest when light transmitted is greatest. This occurs when coloured solution is dilute. The general meter is designed in such a way that it does not show the fraction of light transmitted but it shows the fraction of light absorbed. This is proportional to the concentration of the coloured substance in the test solution.

Thus colorimetric analysis is meant to compare under suitable conditions. The colour produced by a substance in unknown amount with the same colour produced by a known amount of material being determined.

ULTRAVIOLET SPECTROSCOPY

UV region of electromagnetic radiation lies between 200 – 400 nm and visible region lies between 400 – 800 nm. The UV region of the electromagnetic spectrum is subdivided into two spectral regions as follows –

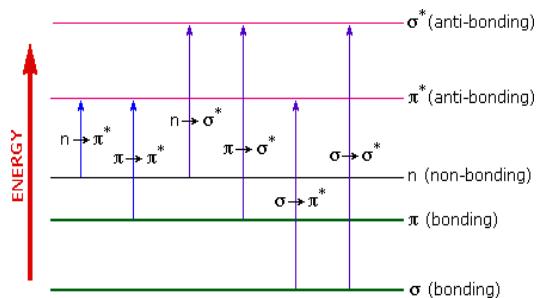
- (i) Near Ultraviolet 200 to 400 nm
 - (ii) Far Ultraviolet 10 to 200 nm
- Or Vacuum Ultraviolet

Ultraviolet absorption spectra originates from transition of electron within a molecule or ion from lower electronic energy level to higher electronic energy level. When an electromagnetic radiation in ultraviolet region is made to pass through a compound containing multiple bonds, it is observed that a part of the incident radiation is usually absorbed which causes electronic excitation. The amount of radiation absorbed depends upon the structure of the compound as well as wavelength of the radiation. The energy of the radiation absorbed causes excitation of electron from lower energy level to higher energy level and the difference of energy is given by – $E = hv$. Thus the actual energy required depends upon the difference in energy between ground state (E_0) and the excited state (E_1) of the electron.

$$E_1 - E_0 = hv$$

UNIT V: INSTRUMENTAL TECHNIQUES IN CHEMICAL ANALYSIS

Electronic Transitions – When UV energy is in the form of quanta the absorption spectrum arising from single electronic transition is expected to consist of a single discrete line, but this does not happen because electronic absorption is super imposed upon Vibrational and rotational sub energy levels. Therefore, the spectra of simple molecules in the gaseous state contain narrow absorption peaks, where in each peak represents a transition from a particular combination of Vibrational and rotational levels in the electronic ground state to a corresponding combination in the excited state. However, in case of polyatomic complex molecules, broad absorption bands are obtained due to coalescence of discrete bands.



When energy is absorbed by a molecule in the UV region, it brings about some changes in the electronic energy of the molecule resulting from transitions of valence electrons. The following three types of electrons are involved in organic molecules.

- (i) **Electrons** – The electrons which are forming double bonds are called reelectrons. These are involved in unsaturated hydrocarbons like trienes and aromatic compounds. In unsaturated systems, π electrons predominantly determine the energy state of electron sheaths which are excited by the absorption of UV or visible light.
- (ii) **Electrons** – The electrons which form single bonds are electrons according to molecular notation. They are involved in saturated bonds between ‘C’ and ‘H’ in paraffins such bonds are also known as ‘S’ bonds. The energy required to excite electrons in bond is very higher than the obtained by UV radiation hence electron do not absorb UV radiation and are not excited by UV radiations.
- (iii) **Electrons** – These are the unshared or non-bonded electrons and are not involved in the bonding between atoms in molecules. Examples – organic compounds containing N, O, or S, halogens. However η – electrons can be excited by UV radiation and hence compounds containing atoms like N, O, S, halogen compounds or unsaturated hydrocarbons may absorb UV radiations.

Representation in the electronic energy levels is as follows:

	Antibonding	σ^*
Energy	Antibonding	π^*
Level	non bonding	η
	Bonding	π
	Bonding	σ^*

Electronic Energy Levels in Simple Organic Molecule

Energy absorbed in the UV region by complex organic molecules results in transitions of valency electrons in the molecules. These transitions are –

- (a) **Transition** – The general characteristics of $\eta \longrightarrow \pi^*$ bonds are (a) They have low intensity (b) They are shifted to shorter wave lengths by more polar solvents as well as electron donating groups. In the spectra of simple molecules $\eta \longrightarrow \pi^*$ transition requires the least energy and the corresponding bonds are of longer wave length. Such transitions are shown by unsaturated compounds which contain atoms like, N, O and S.

These show weaker bands in spectrum. In aldehydes and ketones the band due to $\eta \longrightarrow \pi^*$ transition generally occurs in the range 270-300 nm, while in case of carbonyl compounds, the bands are in the range 300 to 350 nm due to the $\eta \longrightarrow \pi^*$ transitions. This transition is between non-bonding orbital and

UNIT V: INSTRUMENTAL TECHNIQUES IN CHEMICAL ANALYSIS

antibonding orbital. In absorption of uV radiation, the nuclei along with electrons held together plays an important role in determining which wavelength of radiation will be absorbed. The nuclei determine the strength with which electrons are bound and influence the energy gaps between the excited and ground states.

- (b) **π transition** – These types of transitions are related to the transition of electron from bonding orbital to π an antibonding π^* Orbital. In unsaturated molecules, these transitions are possible. Alkenes absorb around 175 nm, alkynes absorb around 170 nm and carbonyl compounds absorb around 188 nm. For example the UV spectrum of ethylene exhibits an intense band at 174 nm and a weak band at 200 nm, both of these are due to $\pi \rightarrow \pi^*$ transitions. According to selection rules only band at 174 nm represent an allowed transition. This can be represented as given in the figure. This is a transition between bonding orbitals and antibonding orbitals.
- (c) **$\eta \rightarrow \sigma^*$ transition** – The energy required for $\eta \rightarrow \sigma^*$ transition is generally less than that required for $\sigma \rightarrow \sigma^*$ transition and their corresponding absorption bands appear at longer wavelengths in the ultraviolet region (180 to 200 nm) saturated compounds with lone pair (non bonding) electrons undergo $\eta \rightarrow \sigma^*$ transitions apart from $\sigma \rightarrow \sigma^*$ transition. In case of saturated molecules which contain atoms having unshared pairs of electrons, $\eta \rightarrow \sigma^*$ transition becomes be determined by commonly available spectrophotometers. For examples – alcohols and amines (containing –OH and –NH₂ groups) absorb between 175 to 200 nm.
- (d) **σ^* transition** – Such transitions occur in case of saturated hydrocarbons which do not contain lone pairs of electrons. The energy required for this type of transitions is very large and absorption band occurs in (126 to 135 nm) for ultraviolet region. For example, methane has λ_{max} at 1219 nm and ethane at 135 nm corresponds spectrophotometers which generally do not operate at wave lengths below 180 nm. Thus this is a transition between bonding orbital and antibonding orbital.

Concept of Chromophore and Auxochrome in the UV spectroscopy

Chromophore- Chromophore is defined as any isolated covalently bonded group that shows a characteristic absorption in the ultraviolet or visible region (200-800 nm). Chromophores can be divided into two groups-

- a) Chromophores which contain p electrons and which undergo pie to pie* transitions. Ethylenes and acetylenes are the example of such chromophores.
- b) Chromophores which contain both p and nonbonding electrons. They undergo two types of transitions; pie to pie* and nonbonding to pie*. Carbonyl, nitriles, azo compounds, nitro compounds etc. are the example of such chromophores.

Auxochrome- An Auxochrome can be defined as any group which does not itself act as a chromophore but whose presence brings about a shift of the absorption band towards the longer wavelength of the spectrum. – OH,-OR,-NH₂,-NHR, -SH etc. are the examples of auxochromic groups.

Absorption and intensity shifts in the UV spectroscopy

There are four types of shifts observed in the UV spectroscopy-

- a) Bathochromic effect- This type of shift is also known as red shift. Bathochromic shift is an effect by virtue of which the absorption maximum is shifted towards the longer wavelength due to the presence of an auxochrome or change in solvents. The nonbonding to pie* transition of carbonyl compounds observes bathochromic or red shift.
- b) Hypsochromic shift- This effect is also known as blue shift. Hypsochromic shift is an effect by virtue of which absorption maximum is shifted towards the shorter wavelength. Generally it is caused due to the removal of conjugation or by changing the polarity of the solvents.
- c) Hyperchromic effect- Hyperchromic shift is an effect by virtue of which absorption maximum increases. The introduction of an auxochrome in the compound generally results in the hyperchromic effect.
- d) Hypochromic effect- Hyperchromic effect is defined as the effect by virtue of intensity of absorption maximum decreases. Hyperchromic effect occurs due to the distortion of the geometry of the molecule with an introduction of new group.

UNIT V: INSTRUMENTAL TECHNIQUES IN CHEMICAL ANALYSIS

INSTRUMENTATION OF UV SPECTROSCOPY

The instrumentation for UV spectroscopy i.e. UV spectrometer is made up of following important components.

- (i) Source of Radiation
- (ii) Monochromator
- (iii) Beam splitter
- (iv) Sample chamber
- (v) Detector
- (vi) Recorder

In UV spectrometer, a beam of light is split into two equal halves, one half of the beam called sample beam is directed through a transparent cell containing a solution of the compound being analyzed and one half (reference beam) is directed through an identical cell that contains only the solvent. The instrument is so designed that it can compare the intensities of the two beams at each wavelength of the region.

Instrumentation and working of the UV spectrometers can be studied simultaneously. Most of the modern UV spectrometers consist of the following parts-

Light Source- Tungsten filament lamps and Hydrogen-Deuterium lamps are most widely used and suitable light source as they cover the whole UV region. Tungsten filament lamps are rich in red radiations; more specifically they emit the radiations of 375 nm, while the intensity of Hydrogen-Deuterium lamps falls below 375 nm.

Monochromator- Monochromator generally composed of prisms and slits. The most of the spectrophotometers are double beam spectrophotometers. The radiation emitted from the primary source is dispersed with the help of rotating prisms. The various wavelengths of the light source which are separated by the prism are then selected by the slits such the rotation of the prism results in a series of continuously increasing wavelength to pass through the slits for recording purpose. The beam selected by the slit is monochromatic and further divided into two beams with the help of another prism.

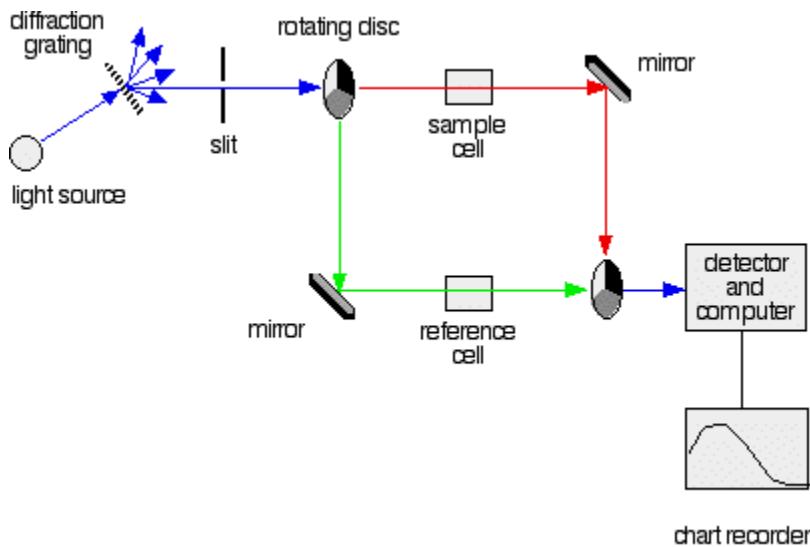
Sample and reference cells- One of the two divided beams is passed through the sample solution and second beam is passé through the reference solution. Both sample and reference solution are contained in the cells. These cells are made of either silica or quartz. Glass can't be used for the cells as it also absorbs light in the UV region.

Detector- Generally two photocells serve the purpose of detector in UV spectroscopy. One of the photocell receives the beam from sample cell and second detector receives the beam from the reference. The intensity of the radiation from the reference cell is stronger than the beam of sample cell. This results in the generation of pulsating or alternating currents in the photocells.

Amplifier- The alternating current generated in the photocells is transferred to the amplifier. The amplifier is coupled to a small servometer. Generally current generated in the photocells is of very low intensity, the main purpose of amplifier is to amplify the signals many times so we can get clear and recordable signals.

Recording devices- Most of the time amplifier is coupled to a pen recorder which is connected to the computer. Computer stores all the data generated and produces the spectrum of the desired compound.

UNIT V: INSTRUMENTAL TECHNIQUES IN CHEMICAL ANALYSIS



APPLICATION OF UV SPECTROSCOPY

- (i) **Qualitative Analysis** – UV spectroscopy is used for characterization of aromatic compounds and conjugated olefins. Identification is done by comparing the UV absorption spectrum of the sample with the UV spectra of known compound for reference.
- (ii) **Detection of Impurities** – Detection of impurities in organic compounds can also be done by UV spectroscopy. (a) If benzene is associated in small quantity with cyclohexane it can be detected by UV spectroscopy by the absorption band of benzene at 255 nm. (b) If impure adiponitrile or hexamethylene diamine is used for the preparation of nylon the product will be of poor quality. The impurities present with the raw materials can be detected by UV method. (c) Purification of organic compounds can be continued until the absorption bands characteristic of the impurities disappear in the spectrum.
- (iii) **Quantitative Analysis** – UV spectroscopy is used for the quantitative analysis of compounds which absorb UV radiation. The determination is carried out on the basis of Beer Lamberts law according to which absorbance is determined by the formula

$$A = \log I_0 / I = \log T = \epsilon lc$$

Where I is molecular absorptivity co-efficient, L is length of the path or cell C is the concentration of solution.

- (iv) **Studying Kinetics of Chemical Reaction** – UV spectroscopy can be used to steady the kinetics of chemical reactions by following the change in concentration of a product or a reactant with time during the reaction.
- (v) **Determination of Dissociation Constants of Weak Acids or Bases** – UV spectroscopy can be used to determine the dissociation constants of acids or bases. The dissociation constant of an acid (HA) is determined by determining the ratio $[H]A/[A]$ spectrometrically from the graph plotted between absorbance and wavelengths at different pH.
- P $K_a = pH + \log [H]A/[A]$
- (vi) **Molecular Weight Determination** – UV Spectroscopy is used in determination of molecular weight of the compound can be converted into a suitable derivative which shows an absorption band in its spectrum for example molecular weight of amine is determined by converting it into picrate. The concentration of amine picrate can be determined by using the formula -

$$C = \frac{\log(I_0 / I_1)}{\epsilon_{\max} \times L}$$

UNIT V: INSTRUMENTAL TECHNIQUES IN CHEMICAL ANALYSIS

- (vii) **Study of Tautomeric Equilibria** – UV spectroscopy can be used to determine the percentage of keto and enol forms present in compounds such as ethyl acethyl acetate by measuring the strength of the respective absorption bands.
- (viii) **Determination of Calcium in Blood Cerum** – Calcium in blood can be indirectly determined by converting the calcium present in 1 ml of cerum as its oxalate, redissolving it into sulphuric acid and treating it with dilute ceric sulphate. The absorption of the excess ceric ion is measured at 315 nm. The amount of calcium in the blood cerum can thus be indirectly calculated.
- (ix) **Determination of Ozone in Environment** – The ozone concentration present in (Smog.) smoke fog in environment can be calculated by measuring its absorption at 260 nm.
- (x) **Detection of functional groups-** UV spectroscopy is used to detect the presence or absence of chromophore in the compound. This technique is not useful for the detection of chromophore in complex compounds. The absence of a band at a particular band can be seen as an evidence for the absence of a particular group. If the spectrum of a compound comes out to be transparent above 200 nm than it confirms the absence of – a) Conjugation b) A carbonyl group c) Benzene or aromatic compound d) Bromo or iodo atoms.
- (xi) **Identification of an unknown compound-** An unknown compound can be identified with the help of UV spectroscopy. The spectrum of unknown compound is compared with the spectrum of a reference compound and if both the spectrums coincide then it confirms the identification of the unknown substance.
- (xii) **Determination of configurations of geometrical isomers-** It is observed that cis-alkenes absorb at different wavelength than the trans-alkenes. The two isomers can be distinguished with each other when one of the isomers has non-coplanar structure due to steric hindrances. The cis-isomer suffers distortion and absorbs at lower wavelength as compared to trans-isomer.

Lambert's and Beer's Law

"When a beam of light is allowed to pass through a transparent medium, the rate of decrease of intensity with the thickness of medium is directly proportional to the intensity of light."

Mathematically, the Lambert's law may be expressed as follows.

$$\begin{aligned} -\frac{dI}{dt} &\propto I \\ -\frac{dI}{dt} &= KI \quad \dots \dots \dots (1) \end{aligned}$$

Where I = intensity of incident light

t = thickness of the medium

K = proportionality constant

Beer's law may be stated as follows:

"Intensity of incident light decreases exponentially as the concentration of absorbing medium increases arithmetically."

The above sentence is very similar to Lambert's law. So,

$$\begin{aligned} I_t &= I_0 e^{-k' c} \\ I_t &= I_0 10^{-0.4343 k' c} \\ I_t &= I_0 10^{K' c} \quad \dots \dots \dots (4) \end{aligned}$$

Where k' and K' = proportionality constants

c = concentration

By combining equation (3) and (4), we get,

UNIT V: INSTRUMENTAL TECHNIQUES IN CHEMICAL ANALYSIS

$$I_t = I_0 10^{-act}$$

$$I_0 / I_t = 10^{act}$$

Where, K and K' = a or ϵ

c = concentration

t or b = thickness of the medium

$$\log I_0 / I_t = \epsilon b c \quad \dots \dots \dots (5)$$

Where ϵ = absorptivity, a constant dependent upon the λ of the incident radiation and nature of absorbing material. The value of ϵ will depend upon the method of expression of concentration.

The ratio I_0 / I_t is termed as transmittance T, and the ratio $\log I_0 / I_t$ is termed as absorbance A. formerly, absorbance was termed as optical density D or extinction coefficient E. the ratio I_0 / I_t is termed as opacity. Thus,

$$A = \log I_0 / I_t \quad \dots \dots \dots (6)$$

From equation (5) and (6),

$$A = \epsilon b c \quad \dots \dots \dots (7)$$

Thus, absorbance is the product of absorptivity, optical path length and the concentration of the solution.

Limitations:

The linearity of the Beer-Lambert law is limited by chemical and instrumental factors. Causes of nonlinearity include:



- deviations in absorptivity coefficients at **high concentrations ($>0.01M$)** due to electrostatic interactions between molecules in close proximity
- **scattering of light due to particulates** in the sample
- **fluorescence or phosphorescence** of the sample
- changes in refractive index at high analyte concentration
- shifts in chemical equilibria as a function of concentration
- non-monochromatic radiation, deviations can be minimized by using a relatively flat part of the absorption spectrum such as the maximum of an absorption band
- stray light

INTRODUCTION

Infrared spectroscopy is one of the most powerful analytical techniques, which provides a spectrum containing a large number of absorption bands from which a wealth of information can be derived about the structure of an organic compound. The absorption of infrared radiation causes the various bands in a molecule to stretch and bend with respect to one another. After absorption of IR radiations, the molecules of chemical substance vibrate at different frequency of vibrations.

When infrared light is passed through the sample, the vibration and the rotational energies of the molecules are increased. Two types of fundamental vibrations are

(1) Stretching vibrations (2) Bending vibrations

(1) Stretching vibrations: In stretching vibrations the atoms move along the bond axis. As a result, the bond length increases or decreases but bond angle remains unchanged. There are two types of stretching vibrations:

(1) Symmetric stretching: In this type the atoms of the molecule move in the same direction.

(2) Asymmetric stretching: In this type the atoms of the molecule move in the opposite direction.

UNIT V: INSTRUMENTAL TECHNIQUES IN CHEMICAL ANALYSIS

(2) Bending or Deformation vibrations: Bending vibrations involves a change in the bond angle whereas the bond length remains unchanged.

There are two types of bending vibrations:

(A) In – plane bending vibrations:

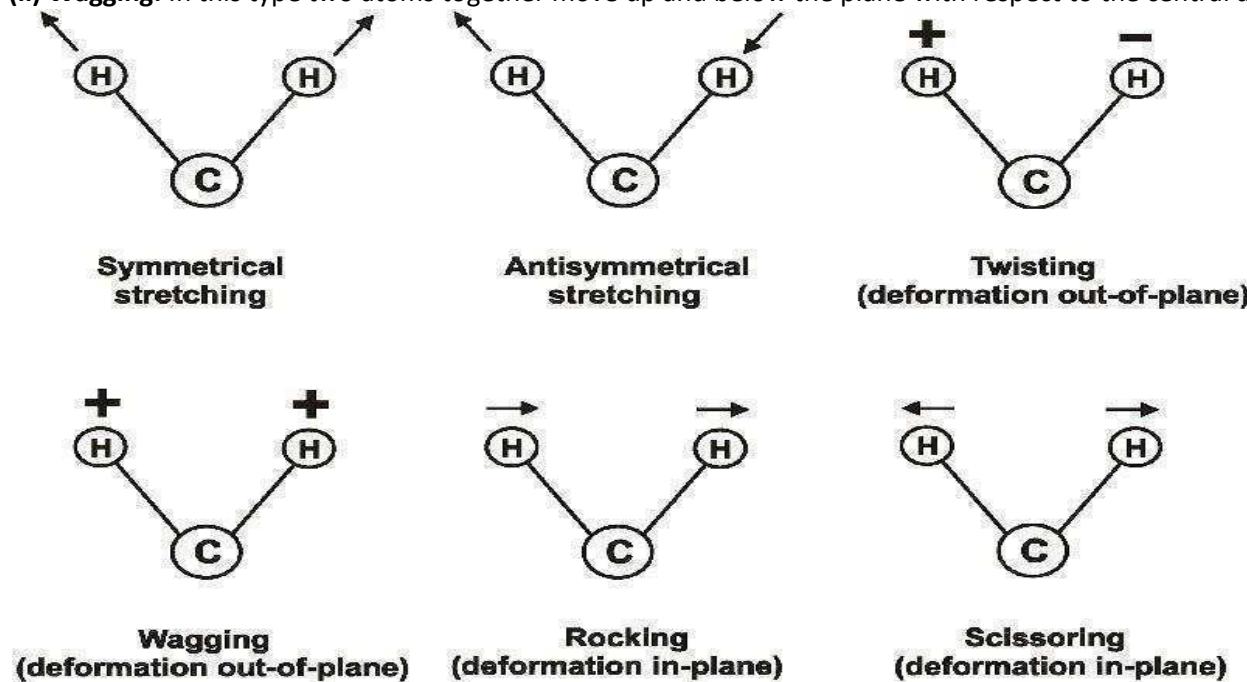
(i) Scissoring: In this type, the atoms move away and come close to each other in the same plane just like the blades of a scissor.

(ii) Rocking: In this type the movement of atoms takes place in the same direction.

(B) Out-of-plane bending vibrations:

(i) Twisting: In this type one atom moves up and the other moves down the plane with respect to the central atom.

(ii) Wagging: In this type two atoms together move up and below the plane with respect to the central atom.



The number of fundamental vibrational modes of a molecule can be calculated are as follows:

A nonlinear molecule containing N atoms has $(3N-6)$ fundamental vibrational modes. For example water is a nonlinear triatomic molecule, therefore, vibrational degrees of freedom of water = $(3N-6)= 3\times 3-6= 3$

So, water is having three fundamental modes of vibration such as symmetrical stretching, asymmetrical stretching and bending vibrations. All the three vibrations are said to be IR active as there is a change in dipole moment during the vibration. So the IR spectrum of water exhibits three absorption bands.

Thus for a vibration to be IR active, there should be a change in dipole moment of the molecule. Homonuclear diatomic molecules like O₂, N₂ and H₂ have zero dipole moments and they are IR inactive.

A linear molecule containing N atoms has $(3N-5)$ fundamental vibrational modes. For example CO₂ is a linear molecule. Therefore, vibrational degrees of freedom of CO₂ = $(3N-5)= 3\times 3-5= 4$

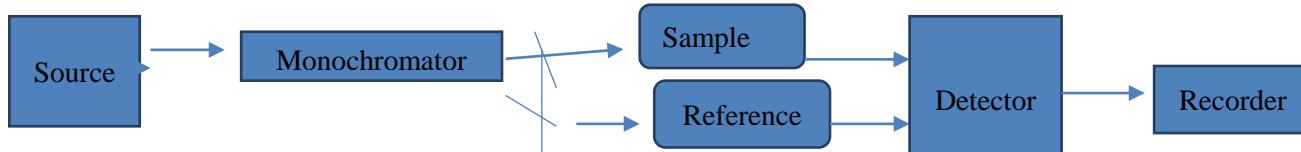
So, Carbon dioxide is having four fundamental modes of vibration such as symmetrical stretching, asymmetrical stretching in plane bending vibrations and out of plane bending vibrations.

UNIT V: INSTRUMENTAL TECHNIQUES IN CHEMICAL ANALYSIS

Instrumentation and working of IR spectrometer:

The essential components of an IR spectrometer are as follows:

- (1) Radiation source (2) Monochromator (3) Sample holder (4) Detector



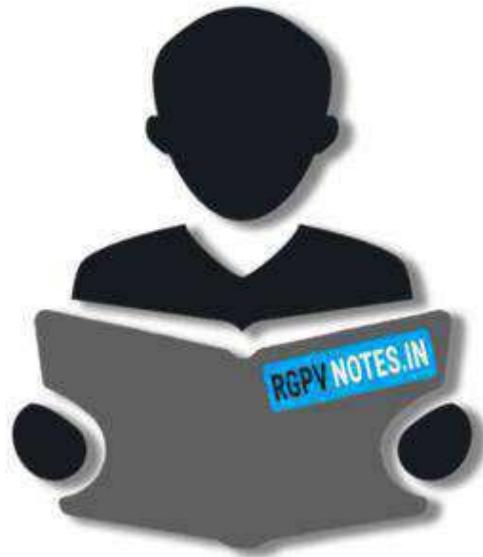
- (1) Radiation source:** The Nernst glower and globar are the most common source of radiation. The Nernst glower consists of a rod of hollow tube about 20 mm long and 1 mm in diameter made by sintering a mixture of oxides of zirconium, thorium and yttrium. Globar is a silicon carbide rod when heated electrically at 1200-2000°C, it glows and produces IR radiations.
- (2) Monochromator:** The radiation source emits radiations of various frequencies. As the sample absorbs only at certain frequencies, it is therefore necessary to select desired frequencies from the radiation source. This has been achieved by monochromators. Prisms and gratings are commonly used for this purpose.
- (3) Sample holder:** The sample holder made up of sodium chloride or potassium bromide. It is used to contain sample solutions as well as reference solution because they are transparent to IR radiation.
- (4) Detector:** The detectors generally convert thermal radiation energy into electrical energy. Thermocouples and bolometer are generally used for this purpose.

Working of IR spectrometer: The radiation from source is passed through monochromator.

Light reflected from the monochromator is diffracted to get the light of specific wavelength. The beam of radiation passed out to monochromator is split into two identical beams, out of which one passes through the reference solution and other through the sample solution. If the frequency of vibration of sample molecules falls within the range of the radiation, the molecule may absorb energy of this frequency from the light. Then the intensity of beam coming out of the sample solution is less than that of beam coming out of the reference solution. If I is the intensity of the sample solution and I_0 that of the reference solution then I/I_0 is called transmittance. A graph is recorded by the instrument which is plot of transmittance versus wave number.

Applications of IR spectroscopy

1. Identification of functional group and structure elucidation.
2. Identification of substances.
3. Studying the progress of the reaction.
4. Detection of impurities.
5. Quantitative analysis



RGPVNOTES.IN

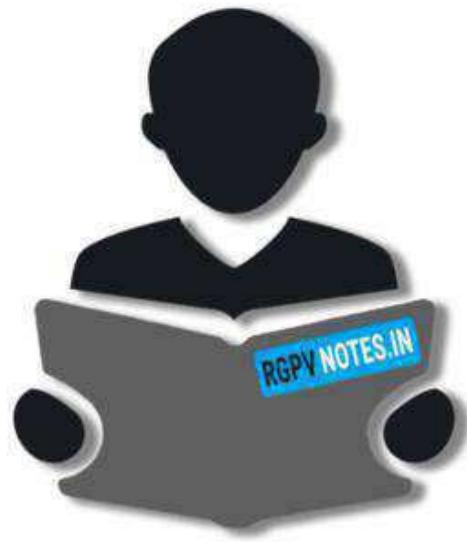
We hope you find these notes useful.

You can get previous year question papers at
<https://qp.rgpvnotes.in> .

If you have any queries or you want to submit your
study notes please write us at
rgpvnotes.in@gmail.com



LIKE & FOLLOW US ON FACEBOOK
facebook.com/rgpvnotes.in



RGPVNOTES.IN

Program : **B.Tech**

Subject Name: **Engineering Chemistry**

Subject Code: **BT-101**

Semester: **1st**



LIKE & FOLLOW US ON FACEBOOK
facebook.com/rgpvnotes.in

PERIODIC TABLE

The periodic table is a list of all known elements arranged in order of increasing atomic number, from 1 to 106. In addition to this, the elements are arranged in such a way that atom with the same number of shells are placed together, and atoms with similar electronic configurations in the outer shell are also placed together. This is achieved as follows:

The elements are arranged in rows and columns. Elements with one shell are placed in the first row (ie H and He), Elements with two shells are placed in the second row (Li to Ne) and so on.

A row of elements thus arranged is called a **period**.

In addition, the elements are aligned vertically (in columns) with other elements in different rows, if they share a similar outer-shell electronic configuration. For example, elements with outer-shell configuration ns^1 are all placed in the same column (Li, Na, K, Rb, Cs, Fr).

A column of elements thus arranged is called a **group**.

According to these principles, the periodic table can be constructed as follows:

I	II																			III	IV	V	VI	VII	O		
																				H						He	
Li	Be																			B	C	N	O	F	Ne		
Na	Mg																			Al	Si	P	S	Cl	Ar		
K	Ca	Sc				Ti	V	Cr	Mn	Fe	Co	Ni	Cu	Zn	Ga	Ge	As	Se	Br	Kr							
Rb	Sr	Y			Zr	Nb	Mo	Tc	Ru	Rh	Pd	Ag	Cd	In	Sn	Sb	Te	I	Xe								
Cs	Ba	La	Ce - Lu		Hf	Ta	W	Re	Os	Ir	Pt	Au	Hg	Tl	Pb	Bi	Po	At	Rn								
Fr	Ra	Ac	Th - Lw																								

Since the electronic configurations of H and He are unusual, they do not fit comfortably into any group. They are thus allocated a group based on similarities in physical and chemical properties with other members of the group.

He is placed in group 0 on this basis, but hydrogen does not behave like any other element and so is placed in a group of its own.

The elements Ce - Lu and Th - Lw belong in the periodic table as shown above. However if they are placed there periods 6 and 7 do not fit onto a page of A4, so they are placed below the other elements in most tables.

All elements belong to one of four main blocks: the s-block, the p-block, the d-block and the f-block.

The s-block elements are all those with only s electrons in the outer shell.

The p-block elements are all those with at least one p-electron in the outer shell.

The d-block elements are all those with at least one d-electron and at least one s-electron but no f or p electrons in the outer shell.

The f-block elements are all those with at least one f-electron and at least one s-electron but no d or p electrons in the outer shell.

I	II				III	IV	V	VI	VII	0									
Li	Be				Ti	V	Cr	Mn	Fe	Co	Ni	Cu	Zn	Ga	Ge	As	Se	Br	Kr
Na	Mg				Zr	Nb	Mo	Tc	Ru	Rh	Pd	Ag	Cd	In	Sn	Sb	Te	I	Xe
K	Ca	Sc			Hf	Ta	W	Re	Os	Ir	Pt	Au	Hg	Tl	Pb	Bi	Po	At	Rn
Rb	Sr	Y	Ce - Lu																
Cs	Ba	La																	
Fr	Ra	Ac	Th - Lw																

Elements colored green are in the s-block

Elements colored blue are in the p-block

Elements colored red are in the d-block

Elements colored black are in the f-block

The physical and chemical properties of elements in the Periodic Table show clear patterns related to the position of each element in the Periodic Table. Elements in the same group show similar properties, and properties change gradually on crossing a Period.

As atomic number increases, the properties of the elements show trends which repeat themselves in each Period of the Periodic Table. These trends are known as **Periodic Trends** and the study of these trends is known as **Periodicity**.

Atomic Radius

The distance from the center of an atom's nucleus to its outermost energy level is its atomic radius. An atom's radius is determined by the size of the electron cloud that surrounds it.

Moving from top to bottom (moving down a group), with each new energy level, the size of the radius increases.

Generally, the radii of atoms decrease in size as you move from left to right on each period. Even though one electron is added to each successive atom (moving left to right on the table), the atomic radius still gets smaller, or "tighter." This happens because of atoms' strong nuclear charge. Each additional proton pulls even tighter to "pull electrons in" closer to the nucleus. Even though the charges are opposite and equal, a proton has almost 2x the attractive power of an electron.

The shell where an electron is added (s, p, d or f) also affects the atomic radius. When electrons are added to the s or p in the outermost shell, the nuclear charge (positive charge from protons in the nucleus) pulls on the outer shell the most.

When electrons are added to a d or f in the inner shell, the outermost electrons have a "shield" from the positive nuclear charge, so they are not as affected as inner electrons.

To summarize:

In general, atomic radius *decreases* from left to right along a period and it *increases* when moving from top to bottom down a group.

INCREASING ATOMIC RADIUS

INCREASING ATOMIC RADIUS

1 H Hydrogen 1.00734																			2 He Helium 4.003	
3 Li Lithium 6.941	4 Be Boron 9.012182																			
11 Na Sodium 22.989770	12 Mg Magnesium 24.3050																			
19 K Potassium 39.0983	20 Ca Calcium 40.078	21 Sc Scandium 44.955910	22 Ti Titanium 47.867	23 V Vanadium 50.9415	24 Cr Chromium 51.9861	25 Mn Manganese 54.93449	26 Fe Iron 55.845	27 Co Cobalt 58.932300	28 Ni Nickel 58.6954	29 Cu Copper 63.546	30 Zn Zinc 65.39	31 Ga Gallium 69.723	32 Ge Germanium 72.61	33 As Arsenic 74.92160	34 Se Selenium 78.96	35 Br Bromine 79.934	36 Kr Krypton 83.810			
37 Rb Rubidium 85.4678	38 Sr Strontium 87.62	39 Y Yttrium 88.90585	40 Zr Zirconium 91.224	41 Nb Niobium 92.91138	42 Mo Molybdenum 95.94	43 Tc Technetium 97.917	44 Ru Ruthenium 101.07	45 Rh Rhodium 102.90550	46 Pd Palladium 106.42	47 Ag Silver 107.8682	48 Cd Cadmium 112.411	49 In Indium 114.813	50 Sn Tin 118.710	51 Te Tellurium 123.760	52 I Iodine 125.60	53 Xe Xenon 126.90447	54 Kr Krypton 131.29			
55 Cs Cesium 132.90345	56 Ba Barium 137.327	57 La Lanthanum 138.90185	72 Hf Hafnium 178.49	73 Ta Tantalum 180.9470	74 W Tungsten 183.84	75 Re Rhenium 186.267	76 Os Osmium 190.23	77 Ir Iridium 192.217	78 Pt Platinum 195.078	79 Au Gold 196.98655	80 Hg Mercury 200.59	81 Tl Thallium 204.3833	82 Pb Lead 207.2	83 Bi Bismuth 210.8	84 Po Polonium (209)	85 At Astatine (210)	86 Rn Radon (222)			
87 Fr Francium (223)	88 Ra Radium (226)	89 Ac Actinium (227)	104 Rf Rutherfordium (261)	105 Db Dubnium (262)	106 Sg Samarium (263)	107 Bh Bohrium (265)	108 Hs Hassium (266)	109 Mt Meitnerium (285)	110 	111 	112 	113 	114 							

Forming Ions

An ion is an atom that has lost or gained electrons.

An atom that loses electrons is positively charged and is called a cation.

An atom that gains extra electrons is negatively charged and is called an anion.

The atomic radius of a cation is smaller than the parent atom's radius, because they have fewer electrons to pull against the positive nuclear charge. With fewer electrons, the positive nuclear charge pulls the remaining electrons tighter, making a smaller atomic radius.

The atomic radius of anions is increased compared to the parent atom. Excess electrons help shield other electrons from the positive nuclear charge. They also act to repel other electrons, pushing them away; this makes the electron cloud "stretch" to a bigger size.

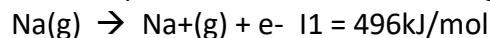
Some atoms easily lose electrons; these are atoms that are unstable until they get rid of some electrons. Other atoms quickly grab extra electrons to become stable.

The ease with which electrons donate or accept electrons (becoming “ionized”) is important for predicting how different elements will react chemically.

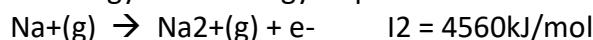
Ionization Energy

“Ionization energy” is the minimum amount of energy required to remove one electron from the ground state (lowest energy state) of the gaseous-state atom.

The first ionization energy is the amount of energy required to remove the first electron from a neutral atom. For example, the first ionization energy of sodium:



Second ionization energy is the energy required to remove the second electron.



The higher the ionization energy, the more difficult it is to remove an electron; in other words, the less likely the atom will give up an electron.

For every atom, ionization energy increases as each electron is removed, because the atom becomes increasingly more positive causing a stronger pull on the remaining electrons.

Ionization energy depends also on atomic radius. When the radius is bigger (and electrons are further away from the nucleus) the ionization energy is lower.

Generally, ionization energy *increases* going left to right along each row and it *decreases* as you go top to bottom down a group. (This is the opposite of the trend for atomic radii)

Factors affecting: -

1. Size of the positive nuclear charge

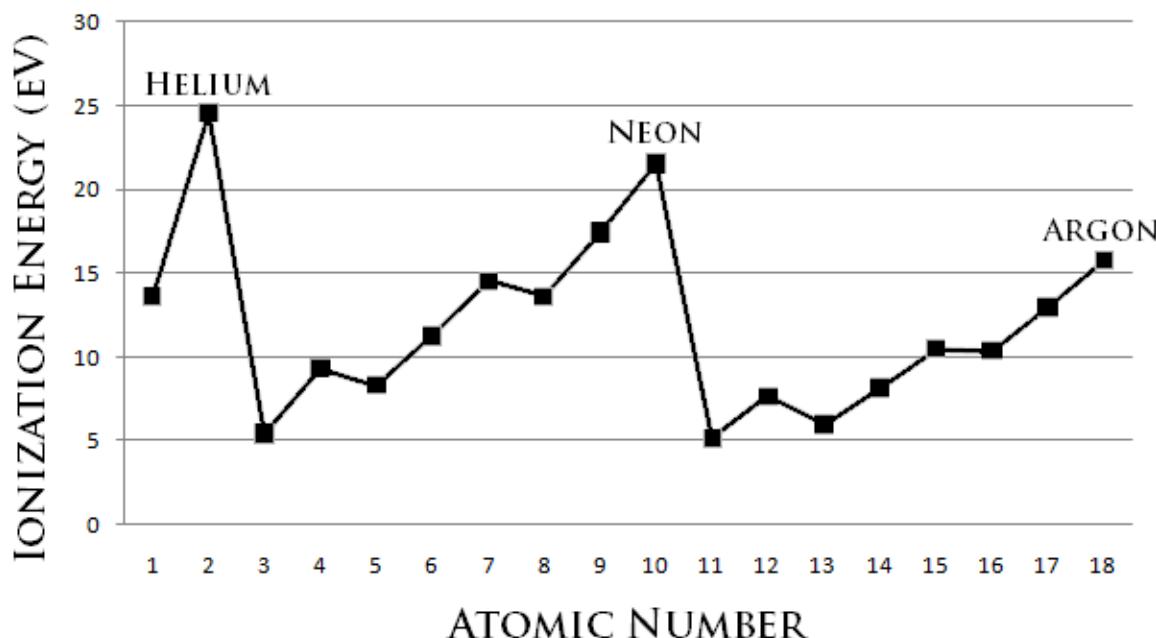
As the nuclear charge increases, its attraction for the outermost electron increases and more energy is required to remove an electron. This means that the ionization energy increases.

2. Size of atom (distance of outermost electron from the nucleus)

As atomic size increases, the attraction of the positive nucleus for the negative electron decreases and less energy is required to remove an electron. This means that the ionization energy decreases.

3. Screening (shielding) effect of inner shell electrons

The outermost electron is screened (shielded) from the attraction of the nucleus by the repelling effect of the inner electrons. As shielding increases, the attraction of the positive nucleus for the negative electron decreases and less energy is required to remove an electron. This means that the ionization energy decreases.



Electron Affinity

Affinity: "a natural liking for, or attraction to a thing."

The energy involved in the addition of an electron to an atom is its electron affinity. This energy is a measure of the attraction (or "affinity") of the atom for the added electron. Usually, when an electron is added (creating an anion), energy is released, so the value is negative.

Example of electron affinity for Chlorine: $\text{Cl(g)} + \text{e}^- \rightarrow \text{Cl}^-(\text{g}) \quad \Delta E = -349 \text{ J}$

In general, electron affinity *increases* from left to right along a period and *decreases* from top to bottom in a group. (Similar to ionization energy and opposite of atomic radii)

Exception: This is NOT true for noble gases .They do not want extra electrons, because that takes more energy.

Remember: Ionization energy is a measure of the energy required to *remove* an electron while electron affinity is a measure of the energy involved in *adding* an electron.

Electro negativity: The tendency for an atom to attract electrons to itself when forming a bond with another element.

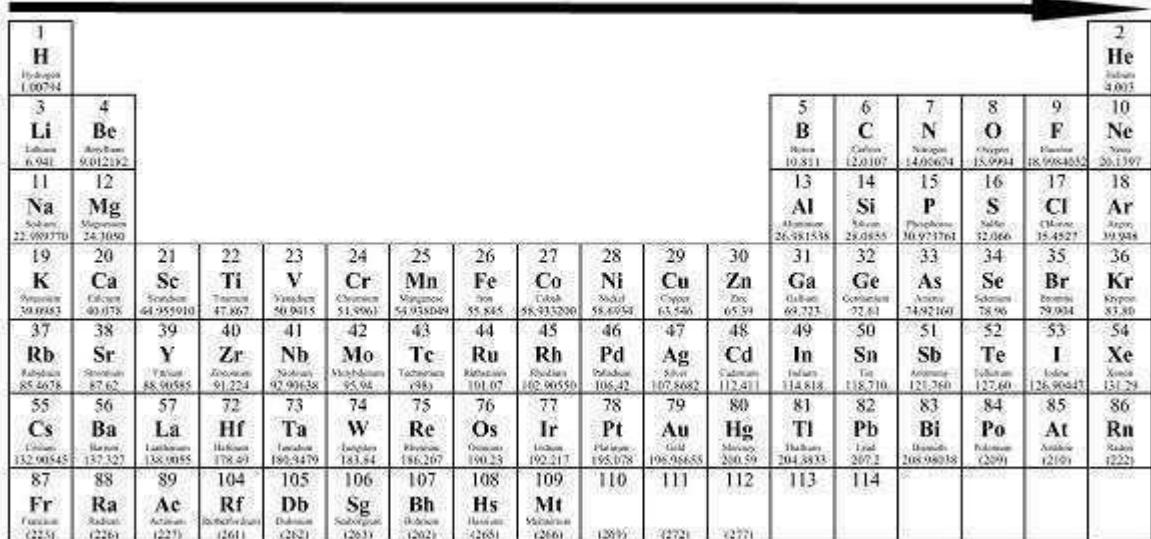
The vertical and horizontal trends for electro negativity (and their explanations) are the SAME as those for ionization energy. There is an exception: fluorine, not helium has the highest electro negativity on the periodic table and similarly the halogens are the group with the highest electro negativity, not the noble gases. This is because the noble gases have a full outer

shell of electrons and do not tend to attract additional electron density. The halogens, on the other hand strongly attract electrons in order to fill their outer shell (remember that they have seven valence electrons, but would be most stable with eight)

- From left to right across a period of elements, electro negativity increases.** If the valence shell of an atom is less than half full, it requires less energy to lose an electron than to gain one. Conversely, if the valence shell is more than half full, it is easier to pull an electron into the valence shell than to donate one.
- From top to bottom down a group, electro negativity decreases.** This is because atomic number increases down a group, and thus there is an increased distance between the valence electrons and nucleus, or a greater atomic radius.
- Important exceptions of the above rules include the noble gases, lanthanides, and actinides.** The noble gases possess a complete valence shell and do not usually attract electrons. The lanthanides and actinides possess more complicated chemistry that does not generally follow any trends. Therefore, noble gases, lanthanides, and actinides do not have electro negativity values.
- As for the transition metals, although they have electro negativity values, there is little variance among them across the period and up and down a group.** This is because their metallic properties affect their ability to attract electrons as easily as the other elements.

According to these two general trends, the *most electronegative element is fluorine*, with 3.98 Pauling units.

INCREASING ELECTRONEGATIVITY



The Periodic Table is shown with arrows indicating the trend of increasing electronegativity. A large horizontal arrow at the top points from left to right, labeled "INCREASING ELECTRONEGATIVITY". Vertical arrows on the left and right sides point downwards, also labeled "INCREASING ELECTRONEGATIVITY". Electronegativity values are provided for each element, generally increasing from left to right across a period and decreasing down a group. The values range from approximately 0.7 for Francium to 3.98 for Fluorine.

Period	Group	Element	Symbol	Electronegativity Value
1	1	Hydrogen	H	1.0074
2	2	Helium	He	4.0073
3	1	Lithium	Li	1.094
3	2	Boron	Be	2.012102
3	3	Sodium	Na	22.089370
3	4	Magnesium	Mg	24.1050
3	5	Aluminum	Al	26.381538
3	6	Silicon	Si	29.0555
3	7	Nitrogen	N	12.0107
3	8	Oxygen	O	14.05634
3	9	Fluorine	F	18.9984832
4	1	Potassium	K	20.00483
4	2	Calcium	Ca	43.075
4	3	Scandium	Sc	44.555910
4	4	Titanium	Ti	47.867
4	5	Vanadium	V	50.9915
4	6	Chromium	Cr	51.9961
4	7	Manganese	Mn	54.935489
4	8	Iron	Fe	55.885
4	9	Cobalt	Co	58.93200
4	10	Nickel	Ni	58.6954
4	11	Zinc	Zn	65.39
4	12	Gallium	Ga	69.723
4	13	Germanium	Ge	72.81
4	14	Arsenic	As	74.92160
4	15	Antimony	Sb	78.96
4	16	Bromine	Br	79.034
4	17	Krypton	Cl	83.938
4	18	Xenon	Ar	89.948
5	1	Rubidium	Rb	85.4678
5	2	Silver	Sr	87.62
5	3	Zincium	Y	88.50585
5	4	Neon	Zr	91.224
5	5	Strontium	Nb	92.99138
5	6	Yttrium	Mo	95.94
5	7	Technetium	Tc	100.07
5	8	Ruthenium	Ru	101.07
5	9	Rhenium	Rh	102.90550
5	10	Palladium	Pd	106.42
5	11	Silver	Ag	107.8682
5	12	Cadmium	Cd	112.411
5	13	Indium	In	114.818
5	14	Tellurium	Sn	118.710
5	15	Antimony	Sb	121.360
5	16	Tellurium	Te	125.60
5	17	Iodine	I	128.90847
5	18	Xenon	Xe	131.29
6	1	Cesium	Cs	132.90545
6	2	Boron	Ba	137.327
6	3	Lanthanum	La	138.9055
6	4	Hafnium	Hf	178.49
6	5	Tantalum	Ta	180.9479
6	6	Tungsten	W	183.84
6	7	Rhenium	Re	186.267
6	8	Osmium	Os	190.23
6	9	Ruthenium	Ir	192.217
6	10	Palladium	Pt	195.178
6	11	Silver	Au	198.95655
6	12	Gold	Hg	200.59
6	13	Hallium	Tl	204.3833
6	14	Thallium	Pb	207.2
6	15	Diodium	Bi	208.98038
6	16	Polonium	Po	229
6	17	Astatine	At	229
6	18	Radon	Rn	222
7	1	Francium	Fr	225
7	2	Radii	Ra	226
7	3	Actinium	Ac	227
7	4	Thorium	Rf	261
7	5	Dubnium	Db	262
7	6	Seaborgium	Sg	263
7	7	Berkelium	Bh	262
7	8	Mendelevium	Hs	265
7	9	Lawrencium	Mt	266
7	10	McDonaldium		269
7	11			272
7	12			277

Figure . Periodic Table showing Electro negativity Trend

Effective nuclear charge – The attractive positive charge of nuclear protons acting on valence electrons.

1. The effective nuclear charge is always less than the total number of protons present in a nucleus due to shielding effect.
2. Effective nuclear charge is behind all other periodic table tendencies.

Shielding effect – the lessening of attractive electrostatic charge difference between nuclear protons and valence electrons by partially or fully filled inner shells.

1. Shielding effect increases with the number of inner shells of electrons.
2. Electrons sharing the same shell do not shield one another from the attractive pull of the nucleus.

Calculating the effective nuclear charge:

An estimate of effective nuclear charge can be obtained from $Z_{\text{eff}} = Z - S$, where Z_{eff} = effective nuclear charge, Z = atomic number, and, S = the screening constant. ""Consider aluminum:
 $[\text{Ne}]3s^23p^1$ "" $Z = 13$ $S = 10$ $Z_{\text{eff}} = Z - S = 13 - 10 = 3+$

Don't forget that Z_{eff} is only an estimate. Actual shielding effect is always greater than the screening constant S because core electrons are much closer to the nucleus than are valence electrons.

Trends

The periodic table tendency for effective nuclear charge:

1. Increase across a period (due to increasing nuclear charge with no accompanying increase in shielding effect).
2. Decrease down a group (although nuclear charge increases down a group, shielding effect more than counters its effect).

ELECTRONIC CONFIGURATION

The **electron configuration** is the distribution of electrons of an atom or molecule (or other physical structure) in atomic or molecular orbital. Electronic configurations describe each electron as moving independently in an orbital, in an average field created by all other orbital.

Atomic number	Symbol	Electron configuration	Atomic number	Symbol	Electron configuration	Atomic number	Symbol	Electron configuration
1	H	$1s^1$	37	Rb	$[Kr]5s^1$	73	Ta	$[Xe]6s^24f^45d^3$
2	He	$1s^2$	38	Sr	$[Kr]5s^2$	74	W	$[Xe]6s^24f^45d^4$
3	Li	$[He]2s^1$	39	Y	$[Kr]5s^24d^1$	75	Re	$[Xe]6s^24f^45d^5$
4	Be	$[He]2s^2$	40	Zr	$[Kr]5s^24d^2$	76	Os	$[Xe]6s^24f^45d^6$
5	B	$[He]2s^22p^1$	41	Nb	$[Kr]5s^14d^4$	77	Ir	$[Xe]6s^24f^45d^7$
6	C	$[He]2s^22p^2$	42	Mo	$[Kr]5s^14d^5$	78	Pt	$[Xe]6s^14f^45d^8$
7	N	$[He]2s^22p^3$	43	Tc	$[Kr]5s^24d^5$	79	Au	$[Xe]6s^14f^45d^{10}$
8	O	$[He]2s^22p^4$	44	Ru	$[Kr]5s^14d^7$	80	Hg	$[Xe]6s^24f^45d^{10}$
9	F	$[He]2s^22p^5$	45	Rh	$[Kr]5s^14d^8$	81	Tl	$[Xe]6s^24f^45d^{10}6p^1$
10	Ne	$[He]2s^22p^6$	46	Pd	$[Kr]4d^{10}$	82	Pb	$[Xe]6s^24f^45d^{10}6p^2$
11	Na	$[Ne]3s^1$	47	Ag	$[Kr]5s^14d^{10}$	83	Bi	$[Xe]6s^24f^45d^{10}6p^3$
12	Mg	$[Ne]3s^2$	48	Cd	$[Kr]5s^24d^{10}$	84	Po	$[Xe]6s^24f^45d^{10}6p^4$
13	Al	$[Ne]3s^23p^1$	49	In	$[Kr]5s^24d^{10}5p^1$	85	At	$[Xe]6s^24f^45d^{10}6p^5$
14	Si	$[Ne]3s^23p^2$	50	Sn	$[Kr]5s^24d^{10}5p^2$	86	Rn	$[Xe]6s^24f^45d^{10}6p^6$
15	P	$[Ne]3s^23p^3$	51	Sb	$[Kr]5s^24d^{10}5p^3$	87	Fr	$[Rn]7s^1$
16	S	$[Ne]3s^23p^4$	52	Te	$[Kr]5s^24d^{10}5p^4$	88	Ra	$[Rn]7s^2$
17	Cl	$[Ne]3s^23p^5$	53	I	$[Kr]5s^24d^{10}5p^5$	89	Ac	$[Rn]7s^26d^1$
18	Ar	$[Ne]3s^23p^6$	54	Xe	$[Kr]5s^24d^{10}5p^6$	90	Th	$[Rn]7s^26d^2$
19	K	$[Ar]4s^1$	55	Cs	$[Xe]6s^1$	91	Pa	$[Rn]7s^25f^66d^1$
20	Ca	$[Ar]4s^2$	56	Ba	$[Xe]6s^2$	92	U	$[Rn]7s^25f^66d^1$
21	Sc	$[Ar]4s^23d^1$	57	La	$[Xe]6s^25d^1$	93	Np	$[Rn]7s^25f^66d^1$
22	Ti	$[Ar]4s^23d^2$	58	Ce	$[Xe]6s^24f^55d^1$	94	Pu	$[Rn]7s^25f^6$
23	V	$[Ar]4s^23d^3$	59	Pr	$[Xe]6s^24f^3$	95	Am	$[Rn]7s^25f^6$
24	Cr	$[Ar]4s^13d^5$	60	Nd	$[Xe]6s^24f^4$	96	Cm	$[Rn]7s^25f^66d^1$
25	Mn	$[Ar]4s^23d^5$	61	Pm	$[Xe]6s^24f^5$	97	Bk	$[Rn]7s^25f^6$
26	Fe	$[Ar]4s^23d^6$	62	Sm	$[Xe]6s^24f^6$	98	Cf	$[Rn]7s^25f^{10}$
27	Co	$[Ar]4s^23d^7$	63	Eu	$[Xe]6s^24f^7$	99	Es	$[Rn]7s^25f^{11}$
28	Ni	$[Ar]4s^23d^8$	64	Gd	$[Xe]6s^24f^55d^1$	100	Fm	$[Rn]7s^25f^{12}$
29	Cu	$[Ar]4s^13d^{10}$	65	Tb	$[Xe]6s^24f^8$	101	Md	$[Rn]7s^25f^{13}$
30	Zn	$[Ar]4s^23d^{10}$	66	Dy	$[Xe]6s^24f^{10}$	102	No	$[Rn]7s^25f^{14}$
31	Ga	$[Ar]4s^23d^{10}4p^1$	67	Ho	$[Xe]6s^24f^{11}$	103	Lr	$[Rn]7s^25f^{14}6d^1$
32	Ge	$[Ar]4s^23d^{10}4p^2$	68	Er	$[Xe]6s^24f^{12}$	104	Rf	$[Rn]7s^25f^{14}6d^2$
33	As	$[Ar]4s^23d^{10}4p^3$	69	Tm	$[Xe]6s^24f^{13}$	105	Db	$[Rn]7s^25f^{14}6d^3$
34	Se	$[Ar]4s^23d^{10}4p^4$	70	Yb	$[Xe]6s^24f^{14}$	106	Sg	$[Rn]7s^25f^{14}6d^4$
35	Br	$[Ar]4s^23d^{10}4p^5$	71	Lu	$[Xe]6s^24f^{14}5d^1$	107	Bh	$[Rn]7s^25f^{14}6d^5$
36	Kr	$[Ar]4s^23d^{10}4p^6$	72	Hf	$[Xe]6s^24f^{14}5d^2$	108	Hs	$[Rn]7s^25f^{14}6d^6$
						109	Mt	$[Rn]7s^25f^{14}6d^7$
						110	Ds	$[Rn]7s^15f^{14}6d^9$
						111	Rg	$[Rn]7s^15f^{14}6d^{10}$

CLASSIFICATION OF ELEMENTS AS s , p , d , f block elements

s block elements :-

The **s-block** is one of four blocks of elements in the periodic table. The element of s- group have a common property. The electron in their most outward electron shell are in the s-orbital. Elements in the s- are in the first two periodic table groups. The elements in group one are called the alkali metals. The elements in group two are called the alkaline earth metals.

Properties:-

- All of the s- elements are metals (except Hydrogen).
- In general, they are shiny, silvery, good conductors of heat and electricity.
- They lose their valence electrons easily. In fact, they lose their trademark s orbital valence electrons so easily that the s- elements are some of the most reactive elements on the periodic table.
- The elements in group 1, known collectively as the alkali metals (except hydrogen), always lose their one valence electron to make a +1 ion. These metals are characterized by being silvery, very soft, not very dense and having low melting points.
- The elements in group 2, known as the alkaline earth metals (except helium), always lose their two valence electrons to make a +2 ion. Like the alkali metals, the alkaline earth metals are silvery, shiny and relatively soft.
- S-elements are famous for being ingredients in fireworks. The ionic forms of potassium, strontium and barium make appearances in firework displays as the brilliant purples, reds and greens.
- The first element in group one, Lithium, and the first in group two, Beryllium, behave differently to other members of their groups. Their behavior is like the second element of the next group. So lithium is similar to magnesium, and beryllium is similar to aluminum.
- In the periodic table this is known as a 'diagonal relationship'. The diagonal relationship is because of similarities in ionic sizes and charge/radius ratio of the element. The similarity between lithium and magnesium is because of their similar sizes Radii, Li=152pm Mg=160pm

Periodic Table of Elements																	
s-Block		p-Block														f-Block	
H		Li Be				B C N O F Ne				Al Si P S Cl Ar				Tl Pb Bi Po At Rn			
Li	Be																
Na	Mg																
K	Ca	Sc	Ti	V	Cr	Mn	Fe	Co	Ni	Cu	Zn						
Rb	Sr	Y	Zr	Nb	Mo	Tc	Ru	Rh	Pd	Ag	Cd						
Cs	Ba	La	Hf	Ta	W	Re	Os	Ir	Pt	Au	Hg						
Fr	Ra	Ac	Rf	Db	Sg	Bh	Hs	Mt	Uun	Uuu	Uub						
* Ce Pr Nd Pm Sm Eu Gd Tb Dy Ho Er Tm Yb Lu																	
** Th Pa U Np Pu Am Cm Bk Cf Es Fm Md No Lr																	

p Block elements : - Elements having a place within the group 13 (i.e. group IIIA) to group 17 (i.e. group VIIA) of the periodic table alongside the group 18 i.e. the zero group elements together frame the p-block of the periodic table.

Position of P Block Elements in the Periodic Table

In the elements of p-block, the last electron enters the furthest p orbital. They have 3 to 8 electrons in the peripheral shell. Consequently, there are six groups of p-block elements in the periodic table numbering from **13 to 18**.

First group: group IIIA called as **Boron group**

Second group: group IVA called as **Carbon group**.

Third group: group VA called as **Nitrogen group**.

Fourth group: group VIA called as **Chalcogens**.

Fifth group: group VIIA called as **Halogens**.

Sixth group: zero group or group 18 called as **Inert or Noble gasses group**.

Characteristics of p block elements: -

- The general valence shell electronic design of p-block elements is $ns\ 2\ np^{1-6}$ (with the exception of He).
- The p-block is the main locale of the periodic table to contain metalloids. The nonmetallic character diminishes down the group though there is a progressive increment in non-metallic character from left to right in the p-block. The metallic character tends to increment down every group while it diminishes as we go from left to right over a period.
- The Atomic Density of elements in p-block increments down the group, this is because of increment in the atom's size down the group. Though it diminishes as we move from left to right over the period, this is because of the lessening in nuclear size of all elements in the p-block over the period.
- The Melting and Boiling points slowly increment down the group in light of the fact that the atomic mass increments down the group and thus the intermolecular forces also increase.

The p-block elements demonstrate a variable oxidation state. The oxidation states increments as we move from left towards right in the periodic table. Boron family (Group 13): - + 3 Carbon family (Group 14): - + 4 Nitrogen family (Group 15): - + 5, Oxygen family (Group 16): - + 6, Halogen family (Group 17): - + 7, Noble gases (Group 18): - + 8

d Block Elements : -

Transition metals are defined as those elements that have (or readily form) partially filled *d* orbitals. The *d*-block elements in groups 3–11 are transition elements. The *f*-block elements, also called *inner transition metals* (the lanthanides and actinides), also meet this criterion because the *d* orbital is partially occupied before the *f* orbitals.

The *d*-block elements are divided into the **first transition series** (the elements Sc through Cu), the **second transition series** (the elements Y through Ag), and the **third transition series** (the element La and the elements Hf through Au). Actinium, Ac, is the first member of the fourth transition series, which also includes Rf through Rg.

CHARACTERISTICS

1. **Metallic nature:** As the number of electrons in the outermost shell is very less i.e. All the transition elements are metals. They show the characteristics of metals such as malleability and ductile in nature and form alloys with several other metals. They also serve as good conductors of heat and electricity. The hard and brittle nature of these elements indicates the presence of covalent bond which is due to the presence of unfilled d-orbital. They are said to form covalent bonding as well as the metallic bonding.
2. **Melting and boiling points:** They show very high melting and boiling points. This can be attributed to the presence of strong metallic bonding due to the overlapping of (n-1) d orbital and covalent bonding of the unpaired d orbital electrons. Since Zn, Cd and Hg have completely filled (n-1)d orbital they are not expected to form covalent bonds. Hence, they show comparatively lower melting point than other d-block elements.
3. **Atomic radii:** A great degree of variation is seen in the atomic radii across each transition series. The atomic radii of the d-block elements within a given series decreases with increase in the atomic number. This is due to the increase in the nuclear charge that attracts the electron cloud inwards resulting in decrease in size. However, the decrease a uniform decrease in atomic radius is not observed across a period. The decrease in atomic radii is small compared to the S and P block elements. This is due to the screening effect caused by the electrons of the (n-1)d sub shell on the outermost shell. The atomic radius increases on descending the group. In a given series, the atomic radius decreases to a minimum for the group VIII elements and then it increases towards the end of the series. This increase in radius towards the end of the series is due to the force of repulsion among the added electrons. A close similarity is observed in the radii of the elements of the second and third transition series due to the filling of 4f subshells.
4. **Ionic radii:** The ionic radius is similar to the pattern of atomic radii. Thus, for ions of a given charge the radius decreases slowly with increase in atomic number.
5. **Atomic volume and Densities:** The atomic volume of transition elements is much lower than those of s and p block elements. This is because of the filling of the (n-1)d orbital that cause an increase in the nuclear charge and pulls the electrons inward. This results in decrease in atomic volume. With the decrease in the atomic volume, the atomic density

these elements increases. Osmium is having a maximum density.

- 6. Ionization potentials:** Transition elements have high ionization energy due to their small size. Their ionization potentials lie between those of s and p block elements. Thus, they are less electropositive than the s-block elements. Hence, they do not form ionic compounds readily like the alkali and alkaline earth metals. They also have the ability to form covalent compounds.

The ionization potentials of d-block elements increase as we move across each series from left to right. However, the increase is not as much as in case of S and P blocks elements. This is due to the screening effect caused by the new electrons that are added into the (n-1) d subshell.

The second ionization energies of the first transition series also increases with the increase in atomic number. However, Cr and Cu are sufficiently higher than those of their neighbors. This is due to their stable electronic configuration.

- 7. Electronic configuration:** The outer electronic configuration remains constant. But, an electron is added to the penultimate shell till the d-sub shell reaches its full capacity. There are three series of elements depending on the n-1 d orbital that is being filled. The orbital are filled in order of their increasing energy i.e. an orbital of lower energy is filled first. Thus 4s orbital with lesser energy is filled first to its full extent then the 3d orbital with higher energy is filled. The exactly half-filled and completely filled d-orbital are extra stable.

The electronic configuration of the first series is given as $1s^2 2s^2 p^6 3s^2 p^6 d^{1-10} 4s^2$

The electronic configuration of the second series is given as $1s^2 2s^2 p^6 3s^2 p^6 d^{1-10} 4s^2 p^6 d^{1-10} 5s^2$

The electronic configuration of the third series is given as $1s^2 2s^2 p^6 3s^2 p^6 d^{1-10} 4s^2 p^6 d^{1-10} 5s^2 p^6 d^{1-10} 6s^2$

Transition elements also show variable oxidation states, tendency to form complexes, magnetic nature and other properties.

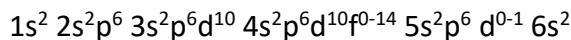
- 8. f Block elements :-** These elements have an incomplete f-orbital of the anti-penultimate shell i.e. inner to the penultimate shell in addition to the incomplete d-orbital of the penultimate shell. F-block elements are also termed as inner transition elements as they are transition elements within the transition elements i.e. d-block elements. An f-orbital can accommodate 14 electrons. This means f-block series can be said to include 14 elements. The general electronic configuration of the f-block elements is given as

$(n-2) f^{1-14}, (n-1) s^2 (n-1) p^6 (n-1) d^{10}, ns^2$. The f-block elements are grouped into two series basing on the nature of the f-orbital of the anti-penultimate shell (4f or 5f) into which the differentiating electron enters.

General Characteristics:

1. Electronic Configuration of Lanthanides: as the 4f and 5d electrons are so close in energy it is not possible to decide whether the electron has entered the 5d or 4f orbital. However, it is considered that the 5d orbital remains vacant and the electrons enter into the 4f orbital except for gadolinium, Gd (Z=64) where the electron enters into the 5d orbital due to the presence of half filled d-orbital. At Ytterbium (z=70) all the 4f orbital's are completely filled and hence, the differentiating electron of the next element that is lutetium (z=71) enters in to the 5d orbital.

2. The complete electronic configuration of Lanthanides can be given as



3. Oxidation States: Lanthanides show variable oxidation states but the degree of variability is less compared to the transition elements. The most stable oxidation state of Lanthanides is +3. In addition to the most stable +2 state, Lanthanides also show +2 and +4 oxidation states. These additional oxidation states also show stability due to presence of either half filled or completely filled or empty 4f subshell. It should be noted that the +2 and +4 oxidation states are unstable in aqueous solutions except for Ce⁺⁴ salts such as the ceric sulphate which acts as an oxidising agent in volumetric analysis.



4. Ionic Radii and Lanthanide contraction: the atomic size decreases with the increase in atomic number as we move across from La to Lu. Thus, among lanthanides, lanthanum has the largest atomic radius and lutetium has the smallest atomic radius. This gradual decrease in the size of an atom is said as lanthanide contraction.

5 Color: All lanthanide metals are silver white. The trivalent lanthanide ions are colored both in solid state and in aqueous solution. It should be noted that the color change is seen only in case of cations. The color of a cation depends on the number of unpaired f electrons. Lanthanides with either half-filled or completely filled orbitals are colorless.

6 Oxidation States: Unlike lanthanides which show the +3 oxidation states, actinides show a variety of oxidation states from +3 to +6. However, +3 and +4 are the principal oxidation states. The +3 oxidation state is the most stable in Ac and all the other elements of the series. The +4 oxidation state is the most stable in Th and Pu. +5 in Pa and Np and +6 is seen in U.

7 **Ionic radius:** In actinides, the ionic radius decreases as we move down the series. This decrease in ionic radius is termed as actinide contraction. This effect is due to poor screening offered by f electrons.

8 **Complex formation:** Unlike lanthanides, they have the tendency to form complexes. Complex formation with ligands such as thio-ethers is also seen.

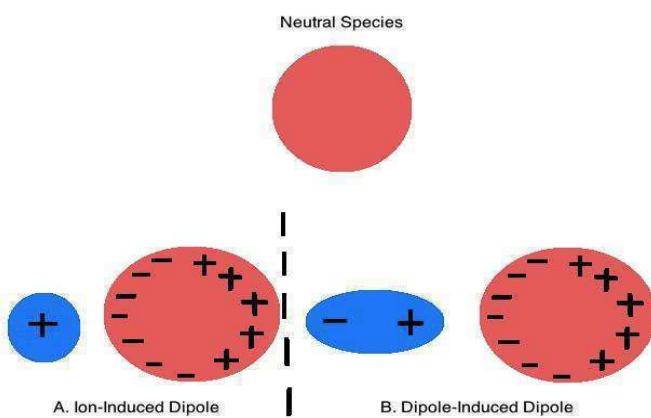
POLARIZABILITY: - **Polarizability** is a measure of how easily an electron cloud is distorted by an electric field. Typically the electron cloud will belong to an **atom** or molecule or ion. The electric field could be caused, for example, by an electrode or a nearby cation or anion.

Neutral non polar species have spherically symmetric arrangements of electrons in their electron clouds. When in the presence of an electric field, their electron clouds can be distorted. The ease of this distortion is defined as the **polarizability** of the atom or molecule. The created distortion of the electron cloud causes the originally non polar molecule or atom to acquire a dipole moment. This induced dipole moment is related to the polarizability of the molecule or atom and the strength of the electric field by the following equation:

$$\mu_{\text{ind}} = \alpha E (1.1) (1.1) \mu_{\text{ind}} = \alpha E$$

Where

- E denotes the strength of the electric field and
- α is the polarizability constant with units of $C m^2 V^{-1}$.



Factors that Influence Polarizability

The relationship between polarizability and the factors of electron density, atomic radii, and molecular orientation is as follows:

1. The greater the number of electrons, the less control the nuclear charge has on charge distribution, and thus the increased polarizability of the atom.
2. The greater the distance of electrons from nuclear charge, the less control the nuclear charge has on the charge distribution, and thus the increased polarizability of the atom.
3. Molecular orientation with respect to an electric field can affect polarizability (labeled Orientation-dependent), except for molecules that are: tetrahedral, octahedral or icosahedral (labeled Orientation-independent). This factor is more important for unsaturated molecules that contain areas of electron dense regions, such as 2,4-hexadiene. Greatest polarizability in these molecules is achieved when the electric field is applied parallel to the molecule rather than perpendicular to the molecule.

Polarizability Influences Dispersion Forces

The dispersion force is the weakest intermolecular force. It is an attractive force that arises from surrounding temporary dipole moments in non polar molecules or species. These temporary dipole moments arise when there are instantaneous deviations in the electron clouds of the non polar species. Surrounding molecules are influenced by these temporary dipole moments and a sort of chain reaction results in which subsequent weak, dipole-induced dipole interactions are created. These cumulative dipole- induced dipole interactions create the attractive dispersion forces. Dispersion forces are the forces that make non polar substances condense to liquids and freeze into solids when the temperature is low enough.

Polarizability affects dispersion forces in the following ways:

- As polarizability *increases*, the dispersion forces also become *stronger*. Thus, molecules attract one another more *strongly* and melting and boiling points of covalent substances *increase with larger molecular mass*.
- Polarizability also affects dispersion forces through the molecular shape of the affected molecules. Elongated molecules have electrons that are easily moved increasing their polarizability and thus strengthening the dispersion forces . In contrast, small, compact, symmetrical molecules are less polarizable resulting in weaker dispersion forces.





RGPVNOTES.IN

We hope you find these notes useful.

You can get previous year question papers at
<https://qp.rgpvnotes.in> .

If you have any queries or you want to submit your
study notes please write us at
rgpvnotes.in@gmail.com



LIKE & FOLLOW US ON FACEBOOK
facebook.com/rgpvnotes.in