UNIT-1

HEATING AND WELDING

Introduction:

Electric heating is extensively used both for domestic and industrial applications. Domestic applications include:

i. Room heaters
ii. Immersion heaters for water heating
iii. Hot plates for cooking
iv. Electric kettles
v. Electric irons
vi. Pop-corn plants
vii. Electric ovens for bakeries
viii. Electric toasters etc.

Industrial applications of electric heating include:

i. Melting of metals
ii. Heat treatment of metals like annealing, tempering, soldering and brazing etc.
iii. Moulding of glass
iv. Baking of insulators
v. Enamelling of copper wires etc.

Advantages of Electric Heating:

Compared to other methods of heating using gas, coal and fire etc., electric heating is far superior for the following reasons:

1. **Cleanliness:** Since neither dust nor ash is produced in electric heating, it is a clean system of heating requiring minimum cost of cleaning. Moreover, the material to be heated does not get contaminated.
2. **No Pollution:** Since no flue gases are produced in electric heating, no provision has to be made for their exit.
3. **Economical:** Electric heating is economical because electric furnaces are cheaper in their initial cost as well as maintenance cost. They do not require big space for installation or for storage of coal and wood. Moreover, there is no need to construct any chimney or to provide extra heat installation.
4. **Ease of Control:** It is easy to control and regulate the temperature of an electric furnace with the help of manual or automatic devices. Temperature can be controlled within ± 5°C which is not possible in any other form of heating.
5. **Special Heating Requirement:** Special heating requirements such as uniform heating of a material or heating one particular portion of the job without affecting its other parts or heating with no oxidation can be met only by electric heating.
6. **Higher Efficiency:** Heat produced electrically does not get wasted through the chimney or any other by-products. Consequently, most of the heat produced is utilised for heating the material itself. Hence, electric heating has higher efficiency as compared to other types of heating.
7. **Better Working Conditions:** Since electric heating produces no irritating noises and the radiation losses are low, it results in low ambient temperature. Hence working with electric furnaces is convenient and cool leading to better working conditions.
8. **Heating of Bad Conductors:** Bad conductors of heat and electricity such as wood, plastic and bakery items can be uniformly and suitably heated with dielectric heating process.

9. **Safety:** Since it responds quickly to the controlled signals, electric heating is quite safe.
10. **Lower Attention and Maintenance Cost:** Electric heating equipment generally will not require much attention and supervision and their maintenance cost is almost negligible. This leads to negligibly small labour charges as compared to other forms of heating.

Different Methods of Heat Transfer:

The different methods by which heat is transferred from a hot body to a cold body are as under:

1. Conduction:

In this mode of heat transfer, one molecule of the body gets heated and transfers some of the heat to the adjacent molecule and so on. There is a temperature gradient between the two ends of the body being heated.

Consider a solid material of cross-section *A* sq.m. and thickness *x* metre as shown in Fig. 1.1. If T_1 and T_2 are the temperatures of the two sides of the slab in °*K*, then heat conducted between the two opposite faces in time *t* seconds is given by:

$$H = \frac{kA(T_1 - T_2)t}{x} \quad \text{MJ}$$

where *k* is the coefficient of the thermal conductivity of the material.

Ex: Refractory heating, the heating of insulating materials, etc.

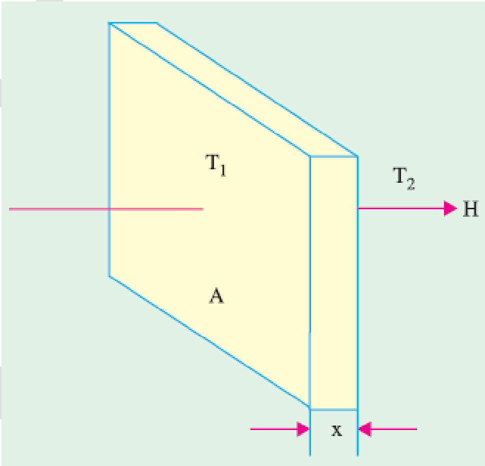


Fig. 1.1

2. Convection:

In this process, heat is transferred by the flow of hot and cold air currents. This process is applied in the heating of water by immersion heater or heating of buildings. The quantity of heat absorbed by the body by convection process depends mainly on the temperature of the heating element above the surroundings and upon the size of the surface of the heater. It also depends, to some extent, on the position of the heater. The amount of heat dissipated is given by

$$H = a(T_1 - T_2)^b \quad \text{W/m}^2$$

Here *a* and *b* are constants whose values are depend upon the heating surface, T_1 and T_2 are the temperatures of the heating element and the fluid in °*C* respectively. In electric furnaces, heat transferred by convection is negligible.

Ex: Immersion water heater.

3. Radiation:

It is the transfer of heat from a hot body to a cold body in a straight line without affecting the intervening medium. The rate of heat emission is given by Stefan's law according to which

Heat dissipated $$H = 5.72 \times ke\left[\left(\frac{T_1}{100}\right)^4 - \left(\frac{T_2}{100}\right)^4\right] \quad \text{W/m}^2$$

or $$H = 5.72 \times 10^4 \times ke\left[\left(\frac{T_1}{1000}\right)^4 - \left(\frac{T_2}{1000}\right)^4\right] \quad \text{W/m}^2$$

where T_1 is the temperature of the source in kelvin, T_2 is the temperature of the substance to be heated in kelvin, *k* is radiating efficiency and *e* is known as emissivity of the heating element.

k = 1, for single element
= 0.5–0.8, for several elements
e = 1, for black body
= 0.9, for resistance heating element

If d is the diameter of the heating wire and l its total length, then its surface area from which heat is radiated = $\pi \times d \times l$. If H is the power radiated per m^2 of the heating surface, then total power radiated as heat = $H \times \pi\ d\ l$. If P is the electrical power input to the heating element, then $\boldsymbol{P = \pi\ d\ l \times H}$.

Ex: Solar heaters.

Methods of Electric Heating:

Basically, heat is produced due to the circulation of current through a resistance. The current may circulate directly due to the application of potential difference or it may be due to induced eddy currents. Similarly, in magnetic materials, hysteresis losses are used to create heat. In dielectric heating, molecular friction is employed for heating the substance. An arc established between an electrode and the material to be heated can be made a source of heat. Bombarding the surface of material by high energy particles can be used to heat the body.

Room heater is a familiar appliance where electric heating is employed.

Fig. 1.2 Different types of Room Heaters

In industries, different types of heatings are adopted as listed below:

a) Power frequency heating utilizing resisting ovens or arc furnace,
b) High frequency heating using induction heating on dielectric heating methods.

Different methods of producing heat for general industrial and domestic purposes may be classified as shown in figure 1.3.

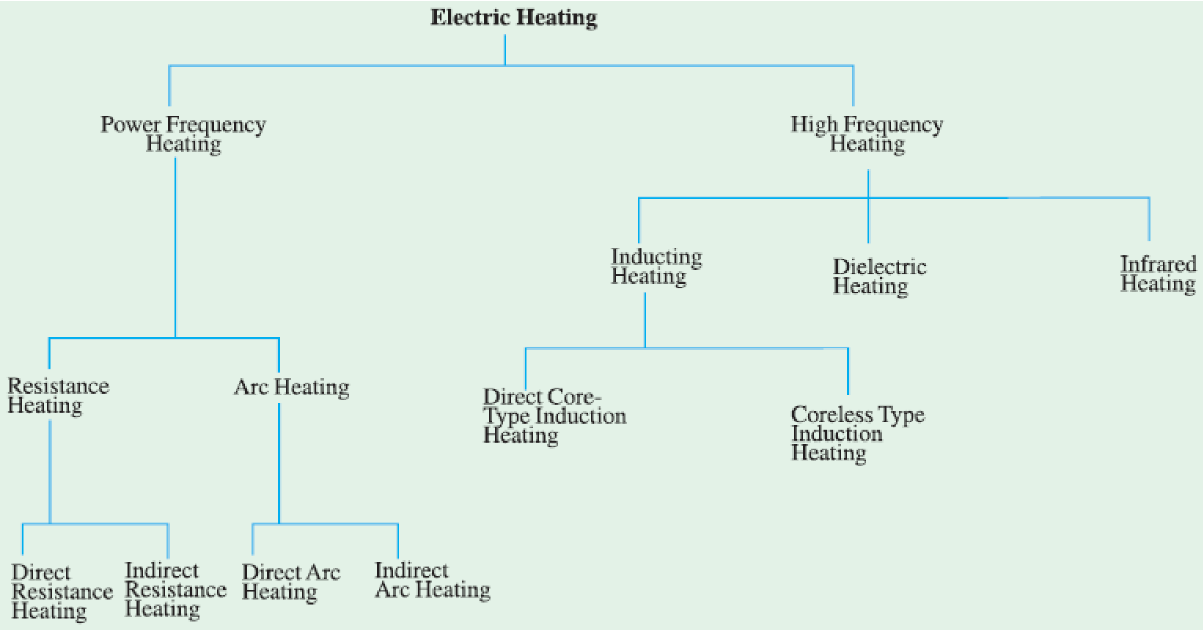


Fig. 1.3 Classification of Electric heating methods

Resistance Heating

It is based on the I^2R effect. When current is passed through a resistance element, I^2R loss takes place which produces heat. There are two methods of resistance heating.

(a) **Direct Resistance Heating:** In this method the material (or charge) to be heated is treated as a resistance and current is passed through it. The charge may be in the form of powder, small solid pieces or liquid. The two electrodes are inserted in the charge and connected to either a.c. or d.c. supply (Fig. 1.4). Obviously, two electrodes will be required in the case of d.c. or single-phase a.c. supply but there would be three electrodes in the case of 3-phase supply. When the charge is in the form of small pieces, a powder of high resistivity material is sprinkled over the surface of the charge to avoid direct short circuit.

(b) **Indirect Resistance Heating:** In this method of heating, electric current is passed through a resistance element which is placed in an electric oven. Heat produced is proportional to I^2R losses in the heating element. The heat so produced is delivered to the charge either by radiation or convection or by a combination of the two. Sometimes, resistance is placed in a cylinder which is surrounded by the charge placed in the jacket as shown in the figure 1.5. This arrangement provides uniform temperature. Moreover, automatic temperature control can also be provided.

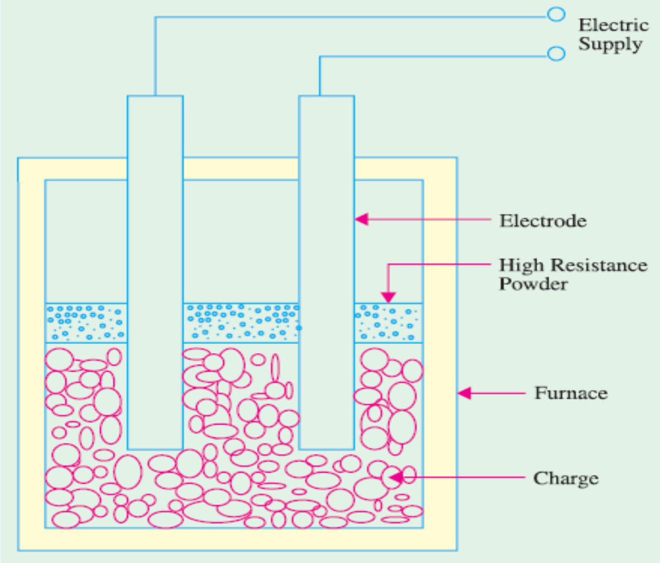


Fig. 1.4 Direct resistance heating

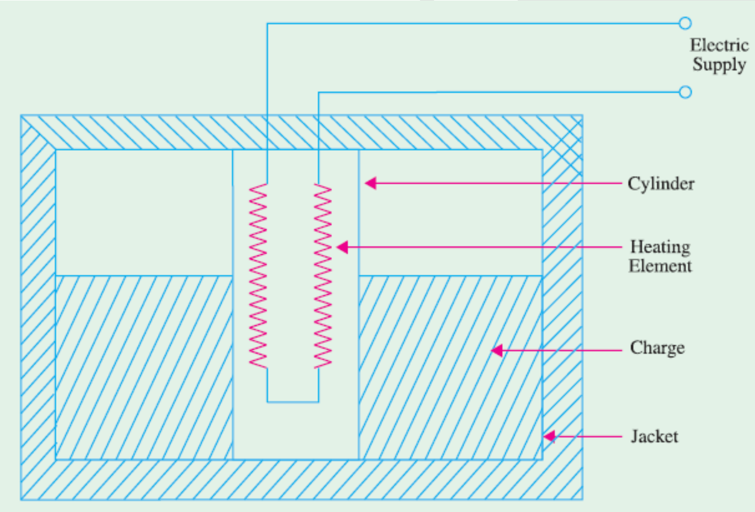


Fig. 1.5 Indirect resistance heating

Requirement of a Good Heating Element:

Indirect resistance furnaces use many different types of heating elements for producing heat. A good heating element should have the following properties:

1) **High Specific Resistance:** When specific resistance of the material of the wire is high, only short length of it will be required for a particular resistance (and hence heat) or for the same length of the wire and the current, heat produced will be more.

2) **High Melting Temperature:** If the melting temperature of the heating element is high, it would be possible to obtain higher operating temperatures.

3) **Low Temperature Coefficient of Resistance:** In case the material has low temperature coefficient of resistance, there would be only small variations in its resistance over its normal range of temperature. Hence, the current drawn by the heating element when cold (*i.e.*, at start) would be practically the same when it is hot.

4) **Oxidising Temperature:** Oxidisation temperature of the heating element should be high in order to ensure longer life.

5) **Positive Temperature Coefficient of Resistance:** If the temperature coefficient of the

resistance of heating element is negative, its resistance will decrease with rise in temperature and it will draw more current which will produce more wattage and hence heat. With more heat, the resistance will decrease further resulting in instability of operation.

6) **Ductile:** Since the material of the heating elements has to have convenient shapes and sizes, it should have high ductility and flexibility.

7) **Mechanical Strength:** The material of the heating element should possess high mechanical strength of its own. Usually, different types of alloys are used to get different operating temperatures. For example maximum working temperature of ***constant*** an (45% Ni, 55% Cu) is 400°C, that of *nichrome* (50%, Ni 20% Cr) is 1150°C, that of ***Kantha*** (70% Fe, 25% Cr, 5% Al) is 1200° C and that of ***silicon carbide*** is 1450°C.

With the passage of time, every heating element breaks open and becomes unserviceable. Some of the factors responsible for its failure are:

1) Formation of hot spots which shine brighter during operation
2) Oxidation
3) Corrosion
4) Mechanical failure

Resistance Furnaces or Ovens

These are suitably-insulated closed chambers with a provision for ventilation and are used for a wide variety of purposes including heat treatment of metals like annealing and hardening etc., stoving of enamelled wares, drying and baking of potteries, vulcanizing and hardening of synthetic materials and for commercial and domestic heating. Temperatures upto 1000°C can be obtained by using heating elements made of nickel, chromium and iron. Ovens using heating elements made of graphite can produce temperatures upto 3000°C. Heating elements may consist of circular wires or rectangular ribbons. The ovens are usually made of a metal framework having an internal lining of fire bricks. The heating element may be located on the top, bottom or sides of the oven. The nature of the insulating material is determined by the maximum temperature required in the oven.

An enclosure for charge which is heated by radiation or convection or both is called a ***heating chamber.***

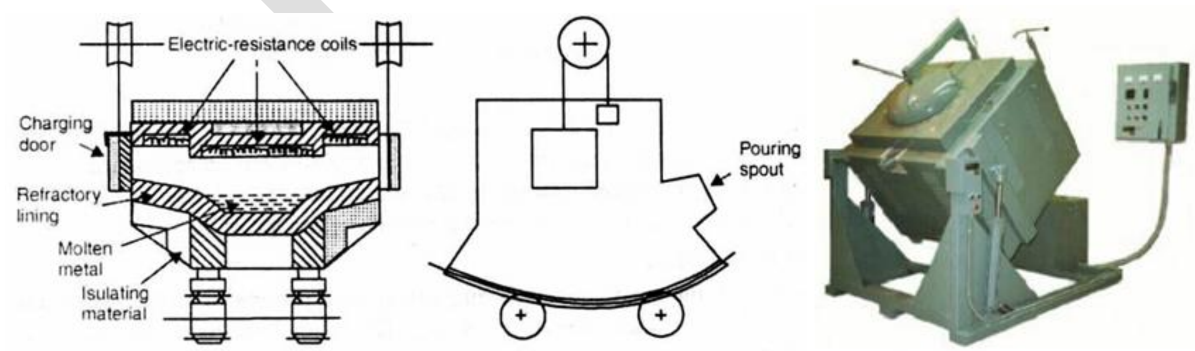


Fig. 1.6 An electric-resistance furnace

A typical resistance furnace is shown in figure 1.6. The solid metal is placed on each of the two inclined hearths and is subjected to heat radiation from the electric-resistance coils located above. When the metal melts, it flows down into a reservoir. The molten metal can be lifted out through the spout (hole) by tilting the whole furnace. Resistance furnace is used mainly for melting aluminium and its alloys and for low melting temperature metals.

Temperature Control of Resistance Furnaces

The temperature of a resistance furnace can be changed by controlling the I^2R or V^2/R losses. Following different methods are used for the above purpose:

1. **Intermittent Switching:** In this case, the furnace voltage is switched ON and OFF intermittently. When the voltage supply is switched off, heat production within the surface is stuck and hence its temperature is reduced. When the supply is restored, heat production starts and the furnace temperature begins to increase. Hence, by this simple method, the furnace temperature can be limited between two limits.

2. **By Changing the Number of Heating Elements:** In this case, the number of heating elements is changed without cutting off the supply to the entire furnace. Smaller the number of heating elements, lesser the heat produced. In the case of a 3-phase circuit, equal number of heating elements is switched off from each phase in order to maintain a balanced load condition.

3. **Variation in Circuit Configuration:** In the case of 3-phase secondary load, the heating elements give less heat when connected in a star than when connected in delta because in the two cases, voltages across the elements is different (Fig.1.7 (a)). In single-phase circuits, series and parallel grouping of the heating elements causes change in power dissipation resulting in change of furnace temperature. As shown in figure 1.7 (b), heat produced is more when all these elements are connected in parallel than when they are connected in series or series-parallel.

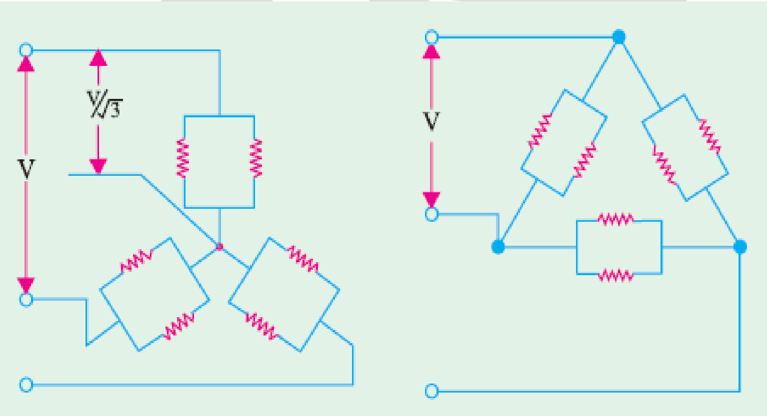


Fig. 1.7 (a)

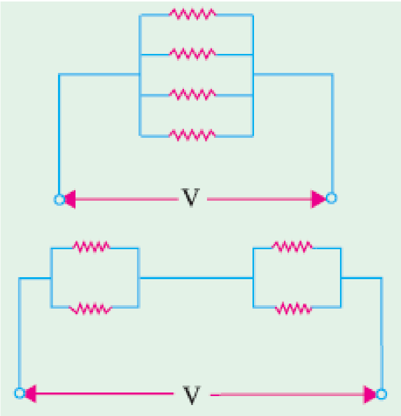


Fig. 1.7 (b)

4. **Change of Applied Voltage: (*a*)** Obviously, lesser the magnitude of the voltage applied to the load, lesser the power dissipated and hence, lesser the temperature produced. In the case of a furnace transformer having high voltage primary, the tapping control is kept in the primary winding because the magnitude of the primary current is less. Consider the multi-tap step-down transformer shown in figure 1.8.

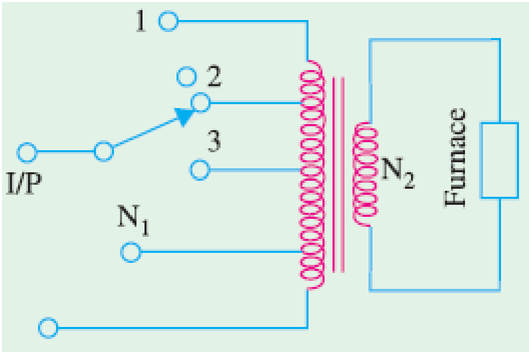


Fig. 1.8

Let the four tappings on the primary winding have 100%, 80%, 60% and 50%. When 100% primary turns are used, secondary voltage is given by $V_2 = (N_2/N_1)V_1$, where V_1 is the input voltage. When 50% tapping is used, the number of primary turns involved is $N_1/2$. Hence, available secondary voltage $V_2 = (2N_2/N_1)V_1$. By selecting a suitable primary tapping, secondary voltage can be increased or decreased causing a change of temperature in the furnace.

(b) Bucking-Boosting the Secondary Voltage. In this method, the transformer secondary is wound in two sections having unequal number of turns. If the two sections are connected in series aiding, the secondary voltage is boosted *i.e.*, increased to ($E_2 + E_3$) as shown in Fig.1.9 (*a*). When the two sections are connected in series-opposing [Fig. 1.9 (*b*)] the secondary voltage is reduced *i.e.*, there is bucking effect. Consequently, furnace voltage becomes ($E_2 - E_3$) and, hence, furnace temperature is reduced.

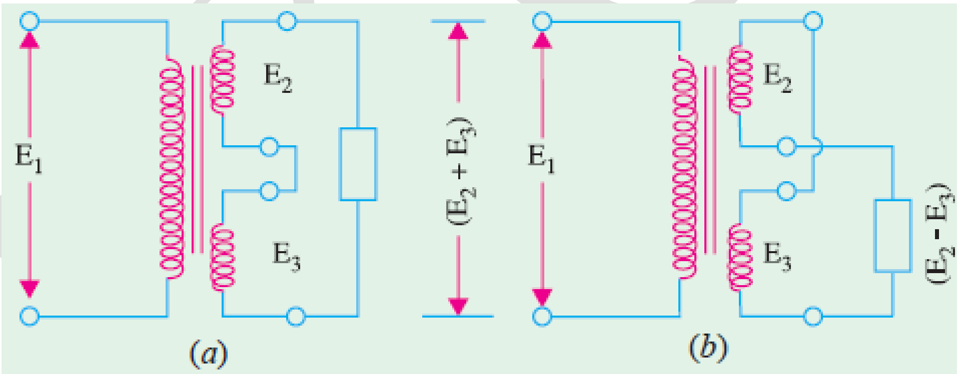


Fig. 1.9

(c) Autotransformer Control. Figure 1.10 shows the use of tapped autotransformer used for decreasing the furnace voltage and, hence, temperature of small electric furnaces. The required voltage can be selected with the help of a voltage selector.

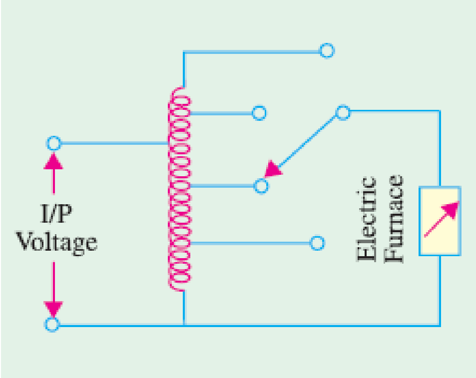


Fig. 1.10

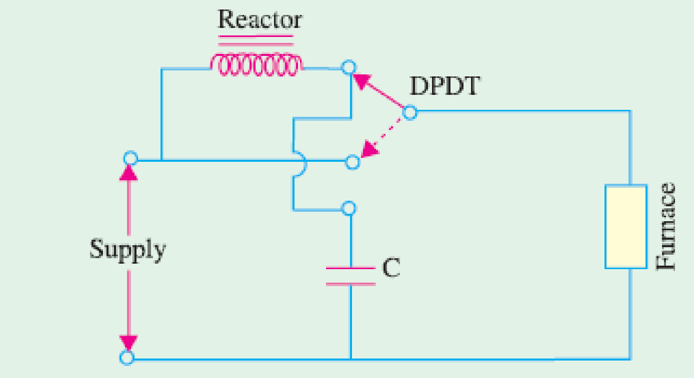


Fig. 1.11

(d) Series Reactor Voltage. In this case, a heavy-duty core-wounded coil is placed in series with the furnace as and when desired. Due to drop in voltage across the impedance of the coil, the voltage available across the furnace is reduced. With the help of D.P.D.T. switch, high/low, two mode temperature control can be obtained as shown in the figure 1.11. Since the addition of series coil reduces the power factor, a power capacitor is simultaneously introduced in the circuit for keeping the p.f. nearly unity. As seen, the inductor is connected in series, whereas the capacitor is in parallel with the furnace.

Design of Heating Elements

Normally, wires of circular cross-section or rectangular conducting ribbons are used as heating elements. The ribbon-type heating element permits the use of higher wattage per unit area compared to the circular-type element. Under steady-state conditions, a heating element dissipates as much heat from its surface as it receives the power from the electric supply. If P is the power input and H is the heat dissipated by radiation, then P = H under steady-state conditions.

Circular-type heating element:

Initially when the heating element is connected to the supply, the temperature goes on increasing and finally reaches high temperature.

Let *V* be the supply voltage of the system and *R* be the resistance of the element, then electric power input,

$$P = \frac{V^2}{R} \quad \text{W}$$

If ρ is the resistivity of the element, *l* is the length, '*a*' is the area, and *d* is the diameter of the element, then:

$$R = \rho\frac{l}{a} = \frac{\rho l}{\frac{\pi d^2}{4}} = \frac{4\rho l}{\pi d^2}$$

Therefore, power input,

$$\boldsymbol{P = \frac{V^2 \pi d^2}{4 \rho l}} \quad \text{W}$$

By rearranging the above equation, we get:

$$\boldsymbol{\frac{l}{d^2} = \frac{\pi V^2}{4 P \rho}}$$

where P is the electrical power input per phase (watt), V is the operating voltage per phase(volts), R is the resistance of the element (Ω), l is the length of the element (m), a is the area of cross-section (m^2), d is the diameter of the element (m), and ρ is the specific resistance (Ω-m).

According to Stefan's law, heat dissipated per unit area is

$$\boldsymbol{H = 5.72 \times ke \left[\left(\frac{T_1}{100} \right)^4 - \left(\frac{T_2}{100} \right)^4 \right]} \quad \text{W/m}^2$$

where T_1 is the absolute temperature of the element (K), T_2 is the absolute temperature of the charge (K), e is the emissivity, and k is the radiating efficiency.

The surface area of the circular heating element:

$$\boldsymbol{S = \pi d l}$$

∴ Total heat dissipated = H × surface area = $H \pi d l$

Under thermal equilibrium,

Power input = heat dissipated

$$\boldsymbol{P = H \times \pi d l}$$

Substituting for P, we get

$$\frac{V^2}{\rho l} \left(\frac{\pi d^2}{4} \right) = H \times \pi d l$$

$$\therefore \boldsymbol{\frac{d}{l^2} = \frac{4 \rho H}{V^2}}$$

By solving the above equations, the length and diameter of the wire can be determined.

Ribbon-type element:

Let ‘w’ be the width and ‘t’ be the thickness of the ribbon-type heating element.

Electrical power input

$$P = \frac{V^2}{R}$$

W.k.t

$$R = \frac{\rho l}{a} = \frac{\rho l}{w \times t}$$

For ribbon or rectangular element, $a = w \times t$

$$\therefore P = \frac{V^2}{\left(\frac{\rho l}{w \times t} \right)}$$

$$\therefore \boldsymbol{\frac{l}{w} = \frac{V^2 t}{P \rho}}$$

The surface area of the rectangular element (S) = 2 $l \times w$

∴ Total heat dissipated = $H \times S = H \times 2\, l\, w$

∴ Under the thermal equilibrium,
Electrical power input = heat dissipated

$\boldsymbol{P = H \times 2\, l\, w}$

$$\boldsymbol{lw = \frac{P}{2H}}$$

By solving the above equations, the length and width of the heating element can be determined.

Example 1.1: A 4.5 kW, 200 V, 1Φ resistance oven is to have nichrome wire heating elements. If the wire temperature is to be 1000°C and that of the charge is 500°C, estimate the diameter and length of the wire. The resistivityy of the nichrome alloy is 42.5 μΩ-m. Assume the radiating efficiency and the emissivity of the element as 1.0 and 0.9 respectively.

Solution:

Given data

Power input (P) = 4.5 kW

Supply voltage (V) = 200 V

Temperature of the source (T_1) = 1,000 + 273 = 1,273 K.

Temperature of the charge T_2= 500 + 273 = 773 K.

According to the Stefan's law,

The amount of heat dissipated ,

$$H = 5.72 \times ke \left[\left(\frac{T_1}{100}\right)^4 - \left(\frac{T_2}{100}\right)^4\right] \quad \mathrm{W/m^2}$$

$$H = 5.72 \times 1 \times 0.9 \left[\left(\frac{1{,}273}{100}\right)^4 - \left(\frac{773}{100}\right)^4\right] \quad \mathrm{W/m^2}$$

$$H = 116.81 \times 10^3 \quad \mathrm{W/m^2}$$

Power, $P = \dfrac{V^2}{R}$

$$P = \frac{V^2}{\left(\frac{\rho l}{a}\right)}$$

$$\left[\text{Since } R = \frac{\rho l}{a}\right]$$

$$P = \frac{V^2 a}{\rho l} = \frac{V^2 \pi d^2}{4\rho l}$$

$$\left[\text{Since the area of circular type element,} \qquad a = \frac{\pi d^2}{4}\right]$$

$$\frac{d^2}{l} = \frac{4P\rho}{V^2\pi} = \frac{4 \times 4.5 \times 10^3 \times 42.5 \times 10^{-6}}{200^2 \times 3.142}$$

$$\frac{d^2}{l} = 6.09 \times 10^{-6} \qquad \ldots\ldots\ldots(i)$$

The heat dissipation is given by:

$P = H \times S = H \times \pi dl$

$$dl = \frac{P}{H\pi} = \frac{4.5 \times 10^3}{116.81 \times 10^3 \times 3.142}$$

$$l = 0.01226/\,d \qquad \ldots\ldots\ldots(ii)$$

Solving (i) and (ii)

***d* = 4.21 mm**

and ***l* = 2.91 m.**

Example 1.2: A 20 kW, 230 V, 1Φ resistance oven employs nickel-chrome strip of 25-mm thick for its heating elements. If the wire temperature is not to exceed 1200°C and the temperature of the charge is to be 700°C. Calculate the width and length of the wire.
Assume the radiating efficiency as 0.6 and emissivity as 0.9. Determine also the temperature of the wire when the charge is cold.

Solution:
Power supplied, $P = 20 \times 10^3$ W.
Let 'w' be the width in meters, t be the thickness in meters, and 'l' be the length in meters.

Then:

Power, $P = \dfrac{V^2}{R} = \dfrac{V^2}{\left(\dfrac{\rho l}{a}\right)} = \dfrac{V^2 \times (wt)}{\rho l}$

[since, $a = wt$]

$$\frac{l}{\mathrm{w}} = \frac{\mathrm{V^2t}}{\mathrm{P\rho}} \quad \text{or} \quad \frac{\mathrm{w}}{l} = \frac{\mathrm{P\rho}}{\mathrm{V^2t}} = \frac{20 \times 10^3 \times 1.016 \times 10^{-6}}{(230)^2 \times 25 \times 10^{-3}}$$

$$\frac{\mathrm{w}}{l} = 1.536 \times 10^{-5} \qquad \ldots\ldots\ldots(i)$$

According to the Stefan's law of heat radiation:

$$H = 5.72 \times ke\left[\left(\frac{T_1}{100}\right)^4 - \left(\frac{T_2}{100}\right)^4\right] \quad \mathrm{W/m^2}$$

$$H = 5.72 \times 0.6 \times 0.9\left[\left(\frac{1200+273}{100}\right)^4 - \left(\frac{700+273}{100}\right)^4\right] \quad \mathrm{W/m^2}$$

$$H = 117.727 \times 10^3 \quad \mathrm{W/m^2}$$

The total heat dissipation is given by:

$P = H \times S = H \times 2lw$ (surface area $S = 2lw$)

$$lw = \frac{P}{2H} = \frac{20 \times 10^3}{2 \times 117.727 \times 10^3}$$

$$lw = 0.0849 \qquad \dots \dots \dots (ii)$$

Solving (i) and (ii)

***w* = 0.2716 mm**

and ***l* = 17.68 m.**

When the charge is cold, it would be at normal temperature, say 25°C.

$$117.727 \times 10^3 = 5.72 \times 0.6 \times 0.9 \left[\left(\frac{T_1}{100}\right)^4 - \left(\frac{25 + 273}{100}\right)^4\right]$$

Solving the above equation we get

T_1=1397.964 K absolute

Or, $\boldsymbol{T_1}$**=1124.96^{0}C**

Example 1.3: Determine the diameter and length of the wire, if a 17-kW, 220-V, 1Φ resistance oven employs nickel-chrome wire for its heating elements. The temperature is not exceeding to 1100°C and the temperature of the charge is to be 500°C. Assume the radiating efficiency as 0.5 and the emissivity as 0.9, respectively.

Solution:

For a circular element:

$$P = \frac{V^2}{R}$$

$$= \frac{V^2}{\frac{\rho l}{A}}$$

$$= \frac{V^2 A}{\rho l}$$

$$= \frac{V^2 \pi d^2}{\rho l 4} \qquad \left[\because \text{The area of circular element } A = \frac{\pi}{4} d^2\right]$$

$$\frac{d^2}{l} = \frac{4\,P\rho}{V^2 \pi}$$

$$= \frac{4 \times 17 \times 10^3 \times 1.016 \times 10^{-6}}{(220)^2 \times 3.14}$$

$$= 4.545 \times 10^{-7}. \qquad (1)$$

According to Stefan's law of heat dissipation:

$$H = 5.72 \times 10^4\, ke \left[\left(\frac{T_1}{1{,}000}\right)^4 - \left(\frac{T_2}{1{,}000}\right)^4\right] \text{W/m}^2$$

$$H = 5.72 \times 10^4 \times 0.5 \times 0.9 \left[\left(\frac{1{,}100 + 273}{1{,}000}\right)^4 - \left(\frac{500 + 273}{1{,}000}\right)^4\right]$$

$$= 82.28 \text{ kW/m}^2.$$

At steady temperature, crucial power input = heat output:

$$P = H \times \pi dl$$

$$dl = \frac{P}{H \times \pi}$$

$$dl = \frac{17 \times 10^3}{82.28 \times 10^3 \times 3.142}$$

$$dl = 0.0658 \qquad \dots\dots\dots\dots\dots (2)$$

Solving Equations (1) and (2), we get:

$$\frac{d^2}{l} \times dl = 4.545 \times 10^{-7} \times 0.0658$$

$$d^3 = 2.99 \times 10^{-8}$$

$$d = 3.1 \text{ mm.}$$

Substitute the value of 'd' in Equation (2) gives:

***l* = 21.226 m**.

Example 1.4: A resistance oven employing nichrome wire is to be operated from 220 V single-phase supply and is to be rated at 16kW. If the temperature of the element is to be limited to 1,170°C and average temperature of the charge is 500°C, find the diameter and length of the element wire. Radiating efficiency = 0.57, Emissivity = 0.9, Specific resistance of nichrome = (109 x10^{-8}) Ω-m.

Solution. $P = 16 \text{ kW} = 16{,}000 \text{ W}$

From Article 47.9, $\frac{l}{d^2} = \frac{\pi V^2}{4\rho P} = \frac{\pi \times (220)^2}{4 \times 109 \times 10^{-8} \times 16{,}000} = 2{,}179{,}660$ ... (*i*)

Now, $H = 5.72 eK\left[\left(\frac{T_1}{100}\right)^4 - \left(\frac{T_2}{100}\right)^4\right] W/m^2 = 5.72 \times 0.9 \times 0.57\left[\left(\frac{1443}{100}\right)^4 - \left(\frac{773}{100}\right)^4\right]$

$= 116{,}752 \text{ W/m}^2$

Now, total heat dissipated/s = electrical power input

$\therefore \quad (\pi d) \times l \times 116{,}752 = 16{,}000; \quad \therefore \quad dl = 0.0436$

or $\quad d^2 l^2 = 0.0019$ ...(*ii*)

From Eqn. (*i*) and (*ii*), $l^3 = 2{,}179{,}660 \times 0.0019 = 4141$

$\therefore \quad l = $ **16.05 m**

$d = 0.0436/16.05 = 2.716 \times 10^{-3} \text{ m} = $ **2.716 mm**

Example 1.5: A 30-kW, 3-ϕ, 400-V resistance oven is to employ nickel-chrome strip 0.254 mm thick for the three star-connected heating elements. If the wire temperature is to be 1,100°C and that of the charge to be 700°C, estimate a suitable width for the strip. Assume emissivity = 0.9 and radiating efficiency to be 0.5 and resistivity of the strip material is 101.6 x10^{-8} Ω-m

Solution. Power/phase = 30 × 1000/3 = 10,000 W, $V_{ph} = 400/\sqrt{3} = 231\text{V}$

If R is the resistance of the strip, $R = V_{ph}^2/P = 231^2/10,000 = 5.34\ \Omega$

Resistance of the strip, $R = \dfrac{\rho l}{wt}$ or $\dfrac{l}{w} = \dfrac{5.34 \times 0.245 \times 10^{-3}}{101.6 \times 10^{-8}} = 1335$ ...(*i*)

Heat dissipated from surface of the strip,

$$H = 5.72 \times 0.9 \times 0.5 \left[\left(\frac{1373}{100} \right)^4 - \left(\frac{973}{100} \right)^4 \right] = 68,400 \text{ W/m}^2$$

Surface area of the strip = $2wl$; Total heat dissipated = $2wl \times H$

∴ $68,400 \times 2 \times wl = 10,000$ or $wl = 0.0731$ ...(*ii*)

From Eqn. (*i*) and (*ii*), we get $w = 0.0731/1335$ or w = **7.4 mm**

Tutorial Problems

A 40kW 3ϕ, 400V resistance oven is to employ Ni-Cr strip of 0.5mm thickness. The heating elements are Y connected. If the wire temperature is to be 1170°C and that of the charge is to be 727°C, estimate the suitable length and width of the wire required. Assume emissivity = 0.6, resistivity of Ni-Cr = 1.03×10^{-6} Ω-m. **[*l* = 10.63m, *w*=5.474mm]**

A 15-kW, 220-V, single-phase resistance oven employs nickel-chrome wire for its heating elements. If the wire temperature is not to exceed 1,000°C and the temperature of the charge is to be 600°C, calculate the diameter and length of the wire. Assume radiating efficiency to be 0.6 and emissivity as 0.9. For nickel chrome resistivity is 1.016×10^{-6} Ω–m. **[3.11 mm, 24.24 m]**

A 30-kW, 3-phase, 400-V resistance oven is to employ nickel-chrome strip 0.025 cm thick for the 3-phase star-connected heating elements. If the wire temperature is to be 1100°C and that of charge is to be 700°C, estimate a suitable width for the strip. Assume radiating efficiency as 0.6 and emissivity as 0.90. The specific resistance of the nichromealloy is 1.03×10^{-6} Ω–m. State any assumptions made. **[6.86 mm]**

Arc Furnaces:

If a sufficiently high voltage is applied across an air-gap, the air becomes ionized and starts conducting in the form of a continuous spark or arc thereby producing intense heat. When electrodes are made of carbon/graphite, the temperature obtained is in the range of 3000°C-3500^{0}C. The high voltage required for striking the arc can be obtained by using a step-up transformer fed from a variable a.c. supply as shown in figure 1.12.

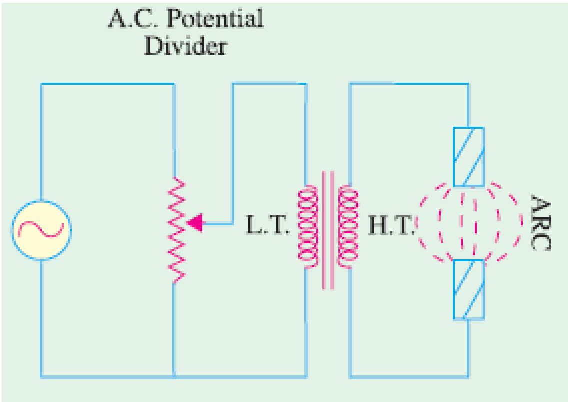


Fig. 1.12

An arc can also be obtained by using low voltage across two electrodes initially in contact with each other as shown in figure 1.13. The low voltage required for this purpose can be obtained by using a step-down transformer. Initially, the low voltage is applied, when the two electrodes are in contact with each other. Next, when the two electrodes are gradually separated from each other, an arc is established between the two.

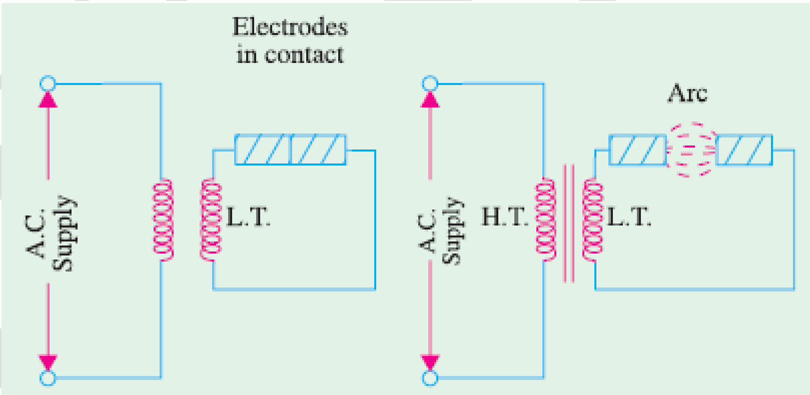


Fig. 1.13

Arc furnaces can be of the following two types :

1. **Direct Arc Furnace**

In this case, arc is formed between the two electrodes and the charge in such a way that electric current passes through the body of the charge as shown in figure 1.14 (*a*). Such furnaces produce very high temperatures.

2. **Indirect Arc Furnace**

In this case, arc is formed between the two electrodes and the heat thus produced is passed on to the charge by radiation as shown in figure 1.14 (*b*).

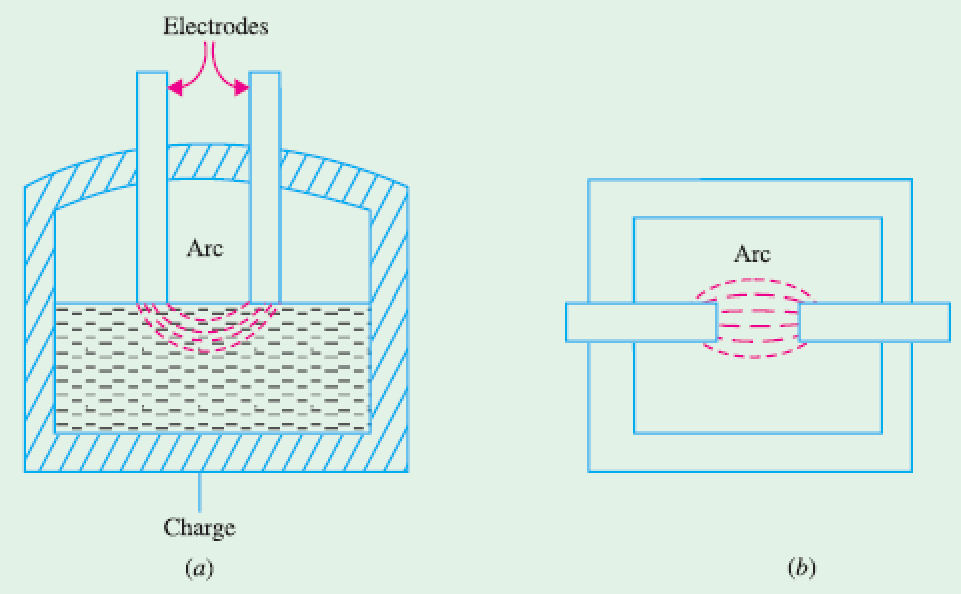


Fig. 1.14

Direct Arc Furnace:

It could be either of conducting-bottom type [Fig. 1.15 (*a*)] or non-conducting bottom type [Fig. 1.15 (*b*)]. As seen from figure 1.15 (*a*), bottom of the furnace forms part of the electric circuit so that current passes through the body of the charge which offers very low resistance. Hence, it is possible to obtain high temperatures in such furnaces. Moreover, it produces uniform heating of charge without stirring it mechanically. In figure 1.15 (*b*), no current passes through the body of the furnace. Most common applications of these furnaces are in the production of steel because of the ease with which the composition of the final product can be controlled during refining.

Most of the furnaces in general use are of non-conducting bottom type due to insulation problem faced in case of conducting bottom.

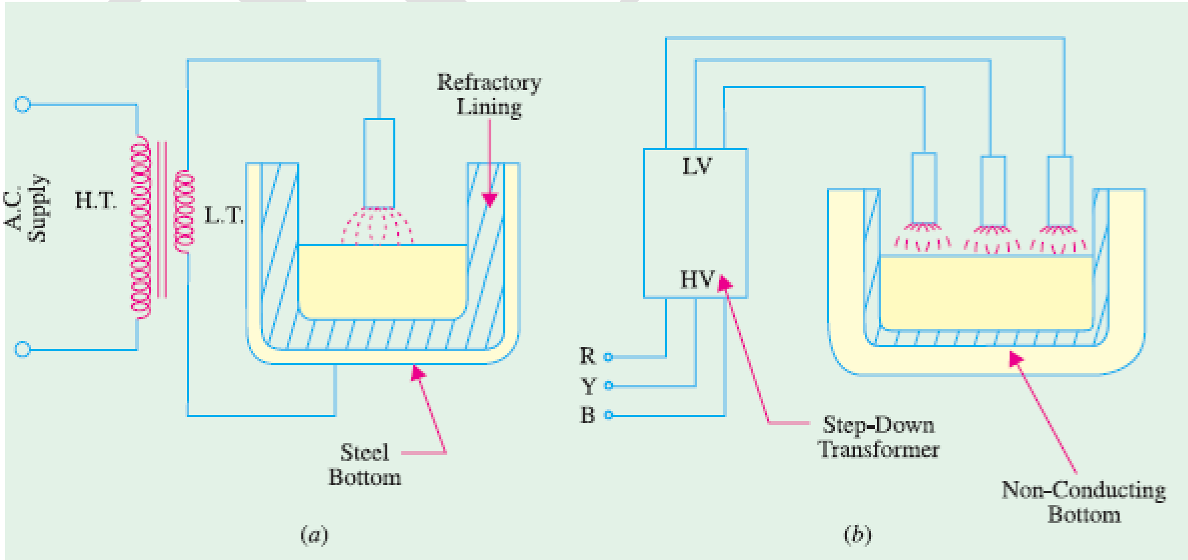


Fig. 1.15

Indirect Arc Furnace:

A single-phase indirect arc furnace which is cylindrical in shape is as shown in figure 1.16. The arc is struck by short-circuiting the electrodes manually or automatically for a moment and then, withdrawing them apart. The heat from the arc and the hot refractory lining is transferred to the top layer of the charge by radiation. The heat from the hot top layer of the charge is further transferred to other parts of the charge by conduction. Since no current passes through the body of the charge, there is no inherent stirring action due to electro-magnetic forces set up by the current. Hence, such furnaces have to be rocked continuously in order to distribute heat uniformly by exposing different layers of the charge to the heat of the arc. An electric motor is used to operate suitable grinders and rollers to impart rocking motion to the furnace. Rocking action provides not only thorough mixing of the charge, it also increases the furnace efficiency in addition to increasing the life of the refractory lining material. Since in this furnace, charge is heated by radiation only, its temperature is lower than that obtainable in a direct arc furnace. Such furnaces are mainly used for melting nonferrous metals although they can be used in iron foundaries where small quantities of iron are required frequently.

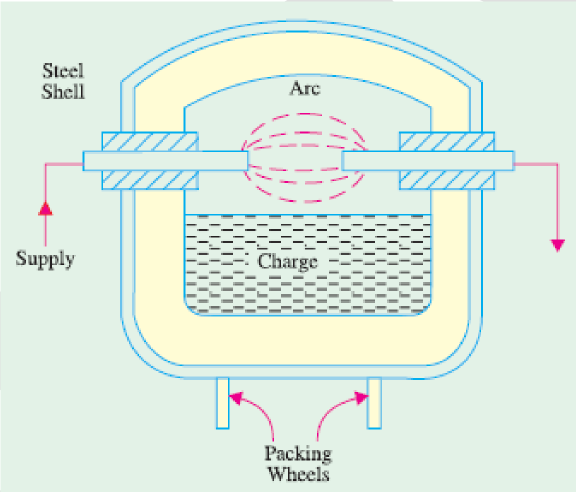


Fig. 1.16

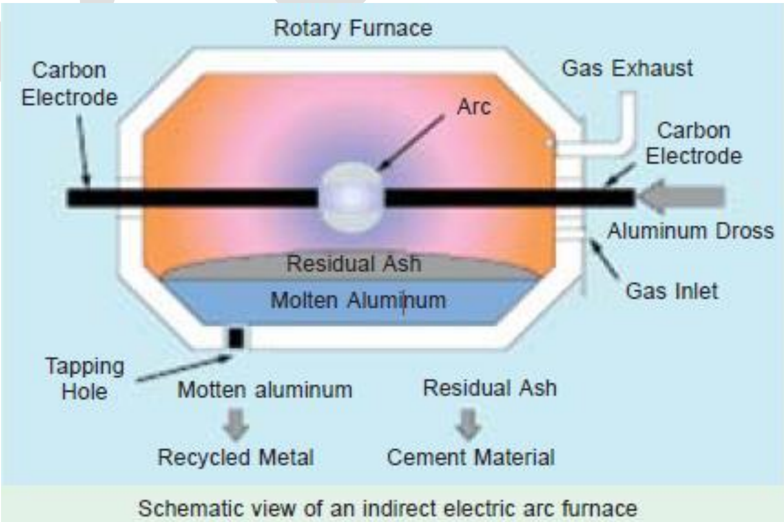


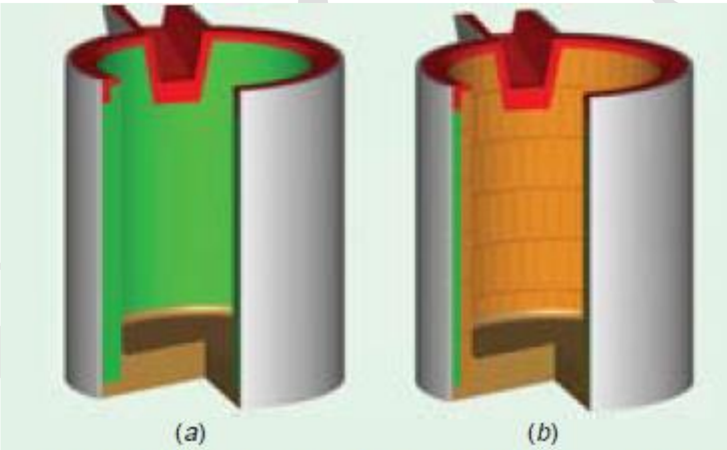
Fig. 1.17

Induction Heating

This heating process makes use of the currents induced by the electro-magnetic action in the charge to be heated. In fact, induction heating is based on the principle of transformer working. The primary winding which is supplied from an a.c. source is magnetically coupled to the charge which acts as a short circuited secondary of single turn. When an a.c. voltage is applied to the primary, it induces voltage in the secondary *i.e.* charge. The secondary current heats up the charge in the same way as any electric current does while passing through a resistance. If V is the voltage induced in the charge and R is the charge resistance, then heat produced = V^2/R. The value of current induced in the charge depends on ***(i)*** magnitude of the primary current ***(ii)*** turn ratio of the transformer ***(iii)*** co-efficient of magnetic coupling. Low-frequency induction furnaces are used for melting and refining of different metals. However, for other processes like case hardening and soldering etc., high-frequency eddy-current heating is employed.

Low frequency induction furnaces employed for the melting of metals are of the following two types:

a) **Core-type Furnaces** — which operate just like a two winding transformer. These can be further sub-divided into ***(i)*** Direct core-type furnaces ***(ii)*** Vertical core-type furnaces and ***(iii)*** Indirect core-type furnaces.
b) **Coreless-type Furnaces** — in which an inductively-heated element is made to transfer heat to the charge by radiation.



Inside views of an induction furnaces: (*a*) Rammed induction furnace and (*b*) bricked induction furnace

Fig. 1.18

Core Type Induction Furnace:

This type of furnace is as shown in figure 1.19. It is essentially a transformer in which the charge to be heated forms a single-turn short-circuited secondary and is magnetically coupled to the primary by an iron core. The furnace consists of a circular hearth which contains the charge to be melted in the form of an annular ring. When there is no molten metal in the ring, the secondary becomes open-circuited there-by cutting off the secondary current. Hence, to start the furnace, molted metal has to be poured in the annular hearth. Since, magnetic coupling between the primary and secondary is very poor, it results in high leakage and low power factor. In order to nullify the effect of increased leakage reactance, low primary frequency of the order of 10 Hz is used. If the transformer secondary current density exceeds 500 A/cm^2 then, due to the interaction of secondary current with the alternating magnetic field, the molten metal is squeezed to the extent that secondary circuit is interrupted. This effect is known as ***pinch effect***.

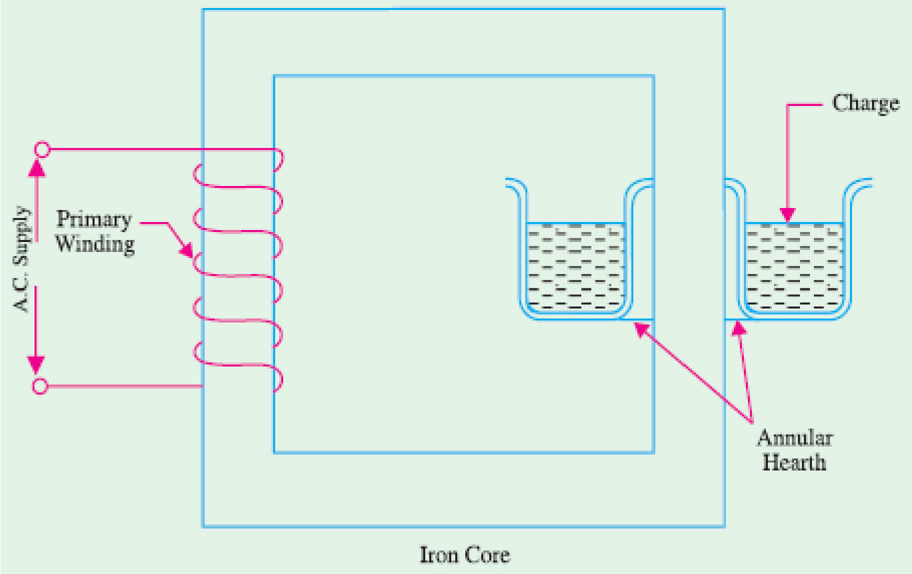


Fig. 1.19 Core Type Induction Furnace

This furnace suffers from the following drawbacks:

1. It has to be run on low-frequency supply which entails extra expenditure on motor-generator set or frequency convertor.
2. It suffers from pinching effect.
3. The crucible for charge is of odd shape and is very inconvenient for tapping the molten charge.
4. It does not function if there is no molten metal in the hearth *i.e.* when the secondary is open. Every time molten metal has to be poured to start the furnace.
5. It is not suitable for intermittent service. However, in this furnace, melting is rapid and clean and temperature can be controlled easily.

Moreover, inherent stirring action of the charge by electro-magnetic forces ensures greater uniformity of the end product.

Vertical Core-Type Induction Furnace (Ajax-Wyatt furnace):

It is also known as Ajax-Wyatt furnace and represents an improvement over the core-type furnace discussed above. As shown in figure 1.20, it has vertical channel (instead of a horizontal one) for the charge, so that the crucible used is also vertical which is convenient from metallurgical point of view. In this furnace, magnetic coupling is comparatively better and power factor is high. Hence, it can be operated from normal frequency supply. The circulation of the molten metal is kept up round the Vee portion by convection currents as shown in Fig. 1.15. As Vee channel is narrow, even a small quantity of charge is sufficient to keep the secondary circuit closed. However, Vee channel must be kept full of charge in order to maintain continuity of secondary circuit.

This fact makes this furnace suitable for continuous operation. The tendency of the secondary circuit to rupture due to pinch-effect is counteracted by the weight of the charge in the crucible. The choice of material for inner lining of the furnace depends on the type of charge used. Clay lining is used for yellow brass. For red brass and bronze, an alloy of magnetia and alumina or corundum is used. The top of the furnace is covered with an insulated cover which can be removed for charging. The furnace can be tilted by the suitable hydraulic arrangement for taking out the molten metal.

This furnace is widely used for melting and refining of brass and other non-ferrous metals. As said earlier, it is suitable for continuous operation. It has a p.f. of 0.8-0.85. With normal supply frequency, its efficiency is about 75% and its standard size varies from 60-300 kW, all single-phase.

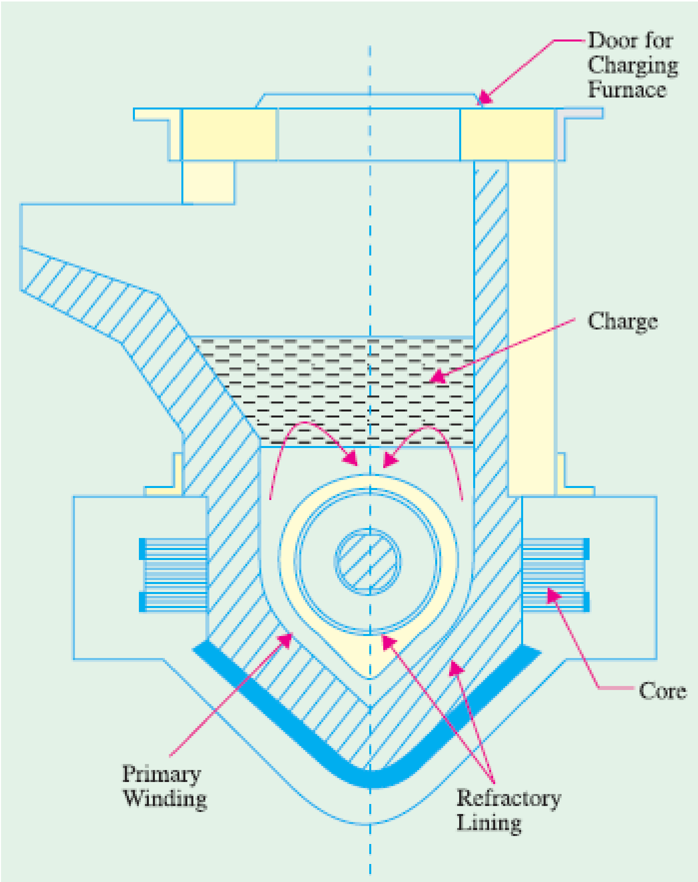


Fig. 1.20 Vertical Core-Type Induction Furnace (Ajax-Wyatt furnace)

Indirect Core-Type Induction Furnace:

In this furnace, a suitable element is heated by induction which, in turn, transfers the heat to the charge by radiation. So far as the charge is concerned, the conditions are similar to those in a resistance oven.

As shown in figure 1.21, the secondary consists of a metal container which forms the walls of the furnace proper. The primary winding is magnetically coupled to this secondary by an iron core. When primary winding is connected to a.c. supply, secondary current is induced in the metal container by transformer action which heats up the container. The metal container transfers this heat to the charge. A special advantage of this furnace is that its temperature can be automatically controlled without the use of an external equipment. The part AB of the magnetic circuit situated inside the oven chamber consists of a special alloy which loses its magnetic properties at a particular temperature but regains them when cooled back to the same temperature. As soon as the chamber attains the critical temperature, reluctance of the magnetic circuit increases manifold thereby cutting off the heat supply. The bar *AB* is detachable and can be replaced by other bars having different critical temperatures.

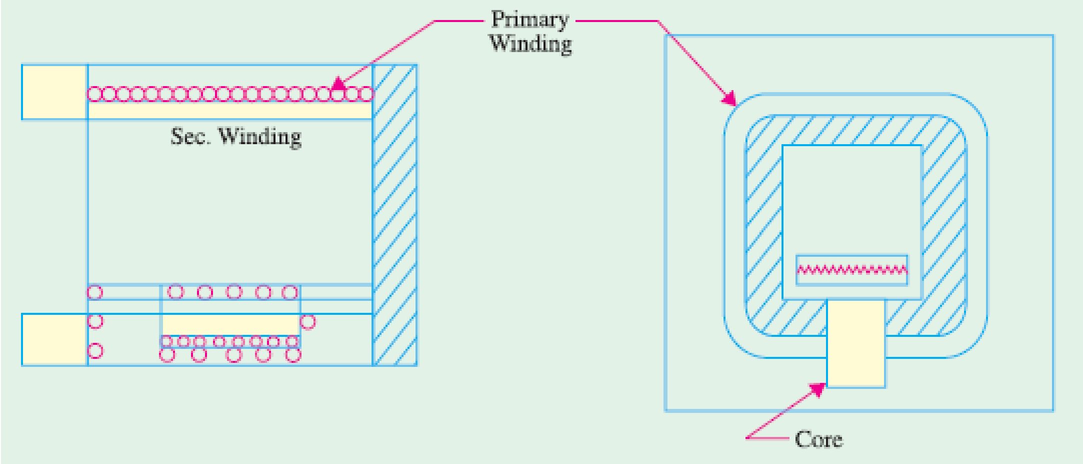


Fig. 1.21

Coreless Induction Furnace:

As shown in figure 1.22, the three main parts of the furnace are ***(i)*** primary coil ***(ii)*** a ceramic crucible containing charge which forms the secondary and ***(iii)*** the frame which includes supports and tilting mechanism. The distinctive feature of this furnace is that it contains no heavy iron core with the result that there is no continuous path for the magnetic flux. The crucible and the coil are relatively light in construction and can be conveniently tilted for pouring.

The charge is put into the crucible and primary winding is connected to a high-frequency a.c. supply. The flux produce by the primary sets up eddy-currents in the charge and heats it up to the melting point. The charge need not be in the molten state at the start as was required by core-type furnaces. The eddy-currents also set up electromotive forces which produce stirring action which is essential for obtaining uniforms quality of metal. Since flux density is low (due to the absence of the magnetic core) high frequency supply has to be used because eddy-current loss is directly proportional to $B^2 f^2$. However, this high frequency increases the resistance of the primary winding due to skin effect, thereby increasing primary Cu losses. Hence, the primary winding is not made of Cu wire but consists of hollow Cu tubes which are cooled by water circulating through them.

Since magnetic coupling between the primary and secondary windings is low, the furnace p.f. lies between 0.1 and 0.3. Hence, static capacitors are invariably used in parallel with the furnace to improve its p.f. Such furnaces are commonly used for steel production and for melting of non-ferrous metals like brass, bronze, copper and aluminium etc., along with various alloys of these elements. Special application of these furnaces include vacuum melting, melting in a controlled atmosphere and melting for precision casting where high frequency induction heating is used. It also finds wide use in electronic industry and in other industrial activities like soldering, brazing hardening and annealing and sterilizing surgical instruments etc.

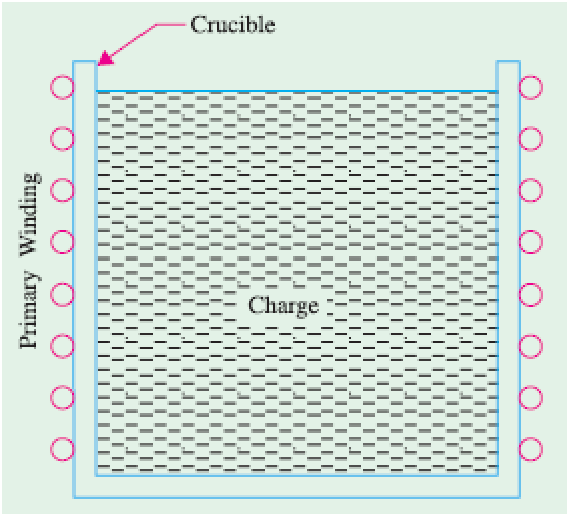


Fig. 1.22

Some of the advantages of coreless induction furnaces are as follows:

1) They are fast in operation.
2) They produce most uniform quality of product.
3) They can be operated intermittently.
4) Their operation is free from smoke, dirt, dust and noises.
5) They can be used for all industrial applications requiring heating and melting.
6) They have low erection and operating costs.
7) Their charging and pouring is simple.

High Frequency Eddy-current Heating

For heating an article by eddy-currents, it is placed inside a high frequency a.c. current-carrying coil (Fig. 1.23). The alternating magnetic field produced by the coil sets up eddy-currents in the article which, consequently, gets heated up. Such a coil is known as heater coil or work coil and the material to be heated is known as ***charge or load***. Primarily, it is the eddy-current loss which is responsible for the production of heat although hysteresis loss also contributes to some extent in the case of non-magnetic materials.

The eddy-current loss $W_e \propto B^2 f^2$. Hence, this loss can be controlled by controlling flux density B and the supply frequency f. This loss is greatest on the surface of the material but decreases as we go deep inside. The depth of the material upto which the eddy-current loss penetrates is given by

$$d = \frac{1}{2\pi}\sqrt{\frac{\rho \times 10^9}{\mu_r . f}}$$

where ρ = resistivity of the molten metal

f = supply frequency and μ_r = relative permiability of the charge

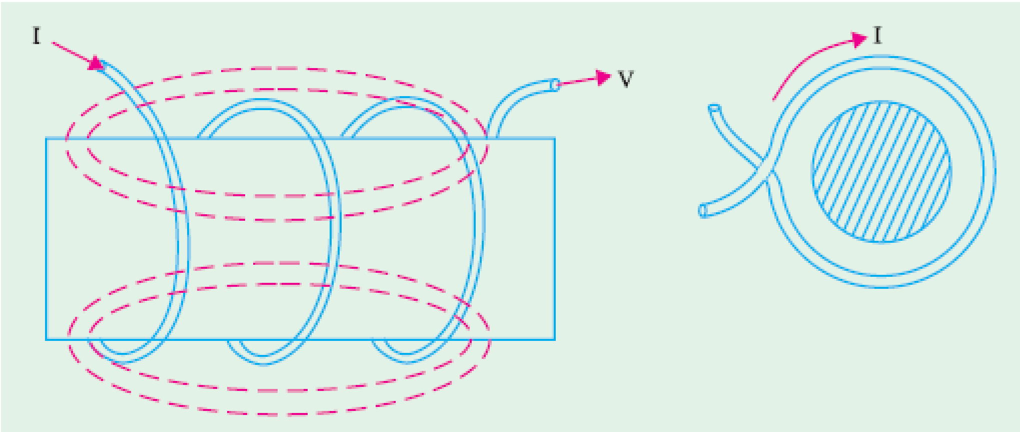


Fig. 1.23

Advantages of Eddy-current Heating

1) There is negligible wastage of heat because the heat is produced in the body to be heated.
2) It can take place in vacuum or other special environs where other types of heating are not possible.
3) Heat can be made to penetrate any depth of the body by selecting proper supply frequently.

Applications of Eddy-current Heating:

1. **Surface Hardening.** The bar whose surface is to be hardened by heat treatment is placed within the working coil which is connected to an a.c. supply of high frequency. The depth upto which the surface is to be hardened can be obtained by the proper selection of frequency of the coil current. After a few seconds, when surface has reached the proper temperature, a.c. supply is cut off and the bar is at once dipped in water.

2. **Annealing.** Normally, annealing process takes long time resulting in scaling of the metal which is undesirable. However, in eddy-current heating, time taken is much less so that no scale formation takes place.

3. **Soldering.** Eddy-current heating is economical for precise high-temperature soldering where silver, copper and their alloys are used as solders.

Example. 1.6. *Determine the efficiency of a high-frequency induction furnace which takes 10 minutes to melt 2 kg of a aluminium initially at a temperature of 20°C. The power drawn by the furnace is 5 kW, specific heat of aluminium = 0.212, melting point of alumimium = 660° C and latent heat of fusion of alumimium. = 77 kcal/kg.*

Solution. Heat required to melt aluminium = 2 × 77 = 154 kcal

Heat required to raise the temperature of aluminium from 20°C to 660°C

$= 2 \times 0.212 \times (660 - 20) = 2 \times 0.212 \times 640 = 271.4$ kcal

Total heat required $= 154 + 271.4 = 425.4$ kcal

Heat required per hour $= 425.4 \times 60/10 = 2552.4$ kcal

Power delivered to aluminium $= 2552.4/860 = 2.96$ kW

$\therefore$ efficiency = output/input $= 2.97/5 =$ **0.594** or **59.4%**

Example 1.7. *A low-frequency induction furnace has a secondary voltage of 20V and takes 600 kW at 0.6 p.f. when the hearth is full. If the secondary voltage is kept constant, determine the power absorbed and the p.f. when the hearth is half-full. Assume that the resistance of the secondary circuit is doubled but the reactance remains the same.*

Solution. Secondary current = $600 \times 10^3/20 \times 0.6 = 5 \times 10^4$ A

If this current is taken as the reference quantity, then secondary voltage is

$$V_2 = 20(0.6 + j\,0.8) \qquad = (12 + j16)\text{V}$$

Hence, secondary impedence, $Z_2 = (12 + j\,16)/5 \times 10^4 = (2.4 + j\,3.2) \times 10^{-4}$ ohm

Now, if the secondary resistance is double, then total impedence when the hearth is half-full is

$$= Z_2 = (4.8 + j3.2) \times 10^{-4} \text{ ohm}$$

Now, secondary current I_2 $= 20/(4.8 + j3.2) \times 10^{-4}$

$$= 20/5.77 \angle 33.7° \times 10^4 = 3.466 \angle -33.7° \times 10^4 \text{ A}$$

Now p.f. = cos 33.7° = 0.832

Hence, power absorbed = $20 \times 3.466 \times 10^4 \times 0.832 \times 10^{-3}$ = **580 kW**

Example 1.8. *A low-frequency induction furnace whose secondary voltage is maintained constant at 10 V, takes 400 kW at 0.6 p.f. when the hearth is full. Assuming the resistance of the secondary circuit to vary inversely as the height of the charge and reactance to remain constant, find the height upto which the hearth should be filled to obtain maximum heat.*

Solution. Secondary current $I_2 = P/V_2 \cos\phi$

$$= 400 \times 10^3/10 \times 0.6 = 6.667 \times 10^4 \text{ A}$$

Impedence of the secondary circuit when hearth is full

$$Z_2 = V_2/I_2 = 10/6.667 \times 10^4 = 1.5 \times 10^{-4}\ \Omega$$

Secondary resistance when hearth is full, $R_2 = Z_2 \cos\phi$

$$= 1.5 \times 10^{-4} \times 0.6 = 0.9 \times 10^{-4}\ \Omega$$

Reactance of the secondary circuit, $X_2 = Z_2 \sin\phi$

$$= 1.5 \times 10^{-4} \times 0.8 = 1.2 \times 10^{-4}\ \Omega$$

In the second, let the height of the charge be x times of the full hearth *i.e.* $h = x$H

Since resistance varies inversely as the height of the charge

$$= R_2 = R_2/x = 0.9 \times 10^{-4}/x\ \Omega$$

Power drawn and hence heat produced will be maximum where secondary resistance equals its reactance.

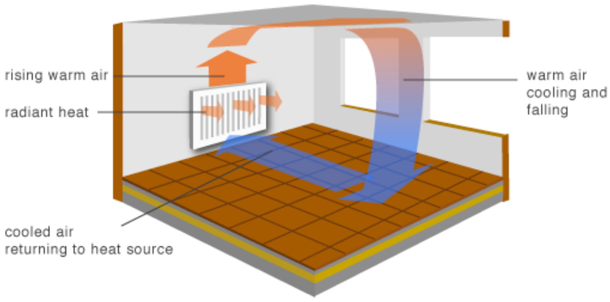
$$\therefore \quad 0.9 \times 10^{-4}/x = 1.2 \times 10^{-4} \quad \text{or} \quad x = \mathbf{3/4}$$

Hence, maximum heat would be produced in the charge when its height is three-fourth the height of the hearth.

Heating of Buildings:

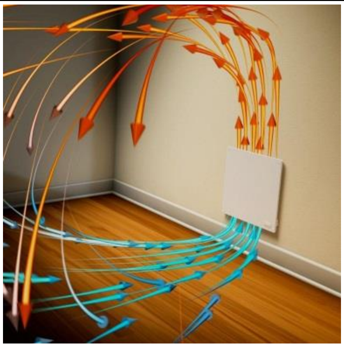
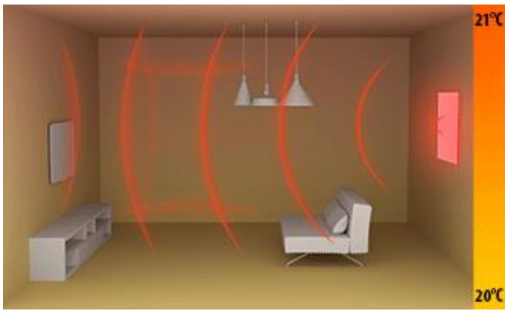
In cold climates or cold weather heating of buildings is undertaken for providing comfort to those occupying buildings. We should therefore have an idea of the heat required by the average human body. The heat dissipation from an average human body is from 100 to 120 watts per hour. Generally the temperature of the human surface is around 27.7^{0}C. The breakup of the heat dissipation from the body is like this: 45% by radiation, 30% by convection and 25% by evaporation. Radiation is dependent upon the body surface temperature and that of the surrounding walls and objects in a room. Heat loss by convection depends upon the temperature of the air in the room, whereas heat loss by evaporation depends upon the humidity of the air. Usually a heating installation aims at providing a temperature of about 18^{0}C which is about 8^{0} to 10^{0}C higher than the temperature of corridors and passages since a higher temperature difference will cause discomfort when a person goes out of a room into the corridor and vice versa. Following are the methods that can be employed for this purpose:

1. ***Radiators:*** the ordinary room heater can heat a small room but where the area to be heated is large, it is not suitable.



2. ***Convectors:*** a spiral of resistance wire is enclosed in a hollow tube. Such tubes are made in length of 3 to 5 meters and are mounted suitably on the walls of the room to be heated. The surface of the tube attains a temperature of about 90^{0}C. The tube may be made in ornamental form also.

3. ***Panel Heating:*** Heating elements are mounted on panels which are fitted on wall or the ceiling of a room. Non-embedded panels have temperature between 40^{0}C to 285^{0}C and the embedded type between 30^{0}C to 40^{0}C.

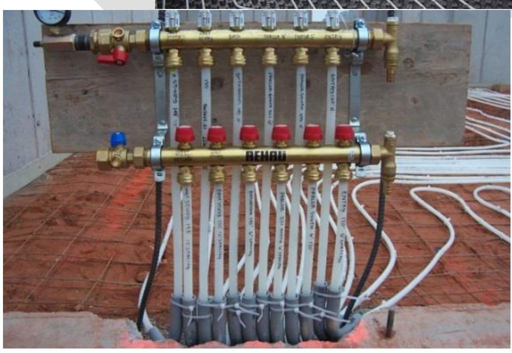
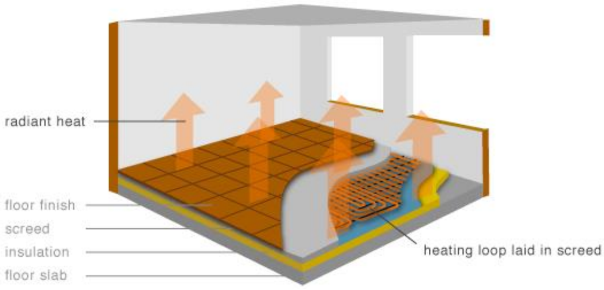




4. ***Hot Water Radiators:*** These are fitted into rooms to be heated and hot water from a central heating system through these.

The Electrode Boiler: the electrode boiler is employed for heating water used radiators. Current is passed through water with the help of electrodes placed in it. AC has to be employed to prevent electrolysis. Cold water has a fairly high resistance so that in order to make it conducting sodium carbonate is added. Electrode boilers operating at 400 V are available in sizes up to 700 kW. For larger sizes, high tension voltage at 11,000 V is used.

The Immersion Heater: when the power required for heating water is less than 200 kW, the electrode boiler is not so economical and in such cases the immersion heater is used. The immersion heater has a resistance in the form of spiral of wire which is suitably insulated by a refractory material. This assembly is placed in a copper or steel tube and properly sealed. Single tubes in ratings of 4 to 5 kW are available. For higher ratings, groups of tubes are used.



A major advantage of the thermal storage system is that water in the storage tank stores enough heat that power is not used during day time, thereby allowing power at low rate to be used during the night.

Dielectric Heating

It is also called high-frequency capacitative heating and is used for heating insulators like wood, plastics and ceramics etc. which cannot be heated easily and uniformly by other methods. The supply frequency required for dielectric heating is between 10-50 MHz and the applied voltage is upto 20 kV. The overall efficiency of dielectric heating is about 50%.

Dielectric Loss:

When a practical capacitor is connected across an a.c. supply, it draws a current which leads the voltage by an angle, which is a little less than 90° or falls short of 90° by an angle δ. It means that there is a certain component of the current which is in phase with the voltage and hence produces some loss called dielectric loss. At the normal supply frequency of 50Hz, this loss is negligibly small but at higher frequencies of 50 MHz or so, this loss becomes so large that it is sufficient to heat the dielectric in which it takes place. The insulating material to be heated is placed between two conducting plates in order to form a parallel-plate capacitor as shown in figure 1.24 (*a*).

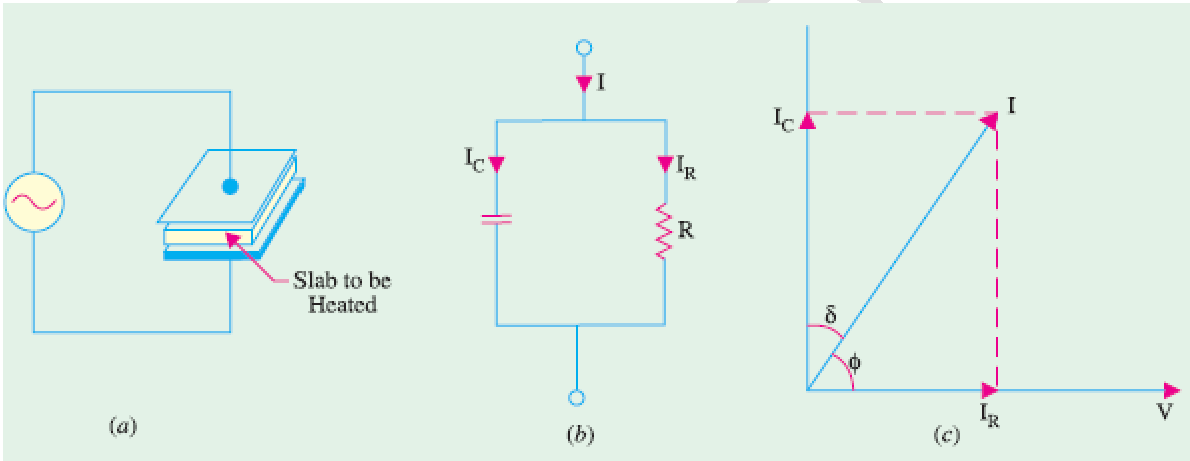


Fig. 1.24

Figure 1.24 (*b*) shows the equivalent circuit of the capacitor and figure 1.24 (*c*) gives its vector diagram.

Power drawn from supply $= VI \cos\phi$

Now, $I_c = I = V/X_c = 2\pi f CV$

$\therefore$ $P = V(2\pi fCV)\cos\phi = 2\pi fCV^2 \cos\phi$

Now, $\phi = (90 - \delta)$, $\cos\phi = \cos(90 - \delta) = \sin\delta = \tan\delta = \delta$

where δ is very small and is expressed in radians.

$P = 2\pi fCV^2\delta$ watts

Here, $C = \epsilon_0 \epsilon_r \frac{A}{d}$ where *d* is the thickness and *A* is the surface area of the dielectric slab.

This power is converted into heat. Since for a given insulator material, *C* and δ are constant, the dielectric loss is directly proportional to $V^2 f$. That is why high-frequency voltage is used in dielectric heating. Generally, a.c. voltage of about 20 kV at a frequency of 10-30 MHz is used.

Advantages of Dielectric Heating:

1. Since heat is generated within the dielectric medium itself, it results in uniform heating.
2. Heating becomes faster with increasing frequency.
3. It is the only method for heating bad conductors of heat.

4. Heating is fastest in this method of heating.
5. Since no naked flame appears in the process, inflammable articles like plastics and wooden products etc. can be heated safely.
6. Heating can be stopped immediately as and when desired.

Applications of Dielectric Heating:

Since cost of dielectric heating is very high, it is employed where other methods are not possible or are too slow. Some of the applications of dielectric heating are as under:

1. For gluing of multilayer plywood boards.
2. For baking of sand cores which are used in the moulding process.
3. For preheating of plastic compounds before sending them to the moulding section.
4. For drying of tobacco after glycerine has been mixed with it for making cigarettes.
5. For baking of biscuits and cakes etc. in bakeries with the help of automatic machines.
6. For electronic sewing of plastic garments like raincoats etc. with the help of cold rollers fed with high-frequency supply.
7. For dehydration of food which is then sealed in air-tight containers.
8. For removal of moistures from oil emulsions.
9. In diathermy for relieving pain in different parts of the human body.
10. For quick drying of glue used for book binding purposes.

Fig. 1.25 Dielectric Heating

Choice of Frequency:

The selection of frequency for heating is important because it has a great bearing on the work to be heated and the method of its heating whether by induction heating or dielectric heating. Furnaces running on power frequency of 50 Hz can be of 1 MW capacity whereas those running on medium frequencies (500 Hz to 1000 Hz) have a capacity of 50 kW and those running on high frequency (1 MHz to 2 MHz) have capacities ranging from 200 kW to 500 kW.

1. **Induction Heating.** While choosing frequency for induction heating, the following factors are considered :

a) Thickness of the surface to be heated. Higher the frequency, thinner the surface that will get heated.

b) The time of continuous heating. Longer the duration of heating, deeper the penetration of heat in the work due to conduction.

c) The temperature to be obtained. Higher the temperature, higher the capacity of the generator required.

2. **Dielectric Heating.** The power consumed during dielectric heating, $P = 2\pi fCV^2 cos\phi$. As seen, P α f × C × V^2 × $cos\phi$. Hence, rate of heat production can be increased by increasing voltage or voltage across any specimen is limited by its thickness or because of the consideration of potential gradient, breakdown voltage and safety etc., Voltages ranging from 600 V to 3000 V are used for dielectric heating, although voltages of 20 kV or so are also used sometimes.

   Rate of heat production can also be increased by applying high potential but it is also limited because of the following considerations:

(a) Possibility of formation of standing waves between the surface of two electrodes having wavelength nearly equal to or more than one quarter of the wavelength of the particular frequency used.

(b) Necessity of employing special matching circuit at higher frequencies due to the fact that maximum power transfer takes place when the oscillator impedance equals the load impedance.

(c) At higher frequencies it is difficult for tuning inductance to resonate with the charge capacitance.

(d) At higher frequencies, it is almost impossible to get uniform voltage distribution.

(e) Since higher frequencies disturb near-by radio station services, special arrangement has to be made to stop radiations from the high-frequency generator used for the purpose.

Infrared Heating:

When tungsten filament lamps are operated at about 2300°C (instead of 3000°C), they produce plenty of heat radiations called ***infrared radiations***. With the help of suitable reflectors, these infrared radiations are focused on the surface to be heated. The lamps so employed have ratings varying from 250 W to 1000 W operating at 115 W. Lower voltage results in robust filaments. With this arrangement, the charge temperature obtain is between 200°C and 300°C. The heat emission intensity obtained is about 7000 W/m^2 as compared to 1500 W/m^2 obtained with ordinary resistance furnaces. In this type of heating, heat absorption remains practically constant whatever the charge temperature whereas it falls rapidly as the temperature of charge rises in the ordinary resistance furnace. Infrared heating is used for paint drying and for drying foundary moulds, for low temperature heating of plastics and for various dehydration and other processes.

Example. 1.9 *. A slab of insulating material 150 cm^2 in area and 1 cm thick is to be heated by dielectric heating. The power required is 400 W at 30 MHz. Material has relative permittivity of 5 and p.f. of 0.05. Determine the necessary voltage. Absolute permittivity = 8.854 × 10^{-12} F/m.*

Solution. $P = 400$ W, p.f. = 0.05, $f = 30 \times 10^6$ Hz

$C = \epsilon_0 \epsilon_r \; A/d = 8.854 \times 10^{-12} \times 5 \times 150 \times 10^{-4}/1 \times 10^{-2} = 66.4 \times 10^{-12}$ F

Now, $P = 2\pi f C V^2 \cos\phi$ or $400 = 2\pi \times 30 \times 10^6 \times 66.4 \times 10^{-12} \times V^2 \times 0.05$ or $V =$ **800 V**

Example. 1.10 *. An insulating material 2 cm thick and 200 cm^2 in area is to be heated by dielectric heating. The material has relative permitivity of 5 and power factor of 0.05. Power required is 400 W and frequency of 40 MHz is to be used. Determine the necessary voltage and the current that will flow through the material. If the voltage were to be limited to 700 V, what will be the frequency to get the same loss?*

Solution. $C = 8.854 \times 10^{-12} \times 5 \times 200 \times 10^{-4}/2 \times 10^{-2} = 44.27 \times 10^{-12}$ F

$$P = 2\pi f C V^2 \cos\phi \quad \text{or} \quad V = \sqrt{400/2\pi \times 40 \times 10^6 \times 44.27 \times 10^{-12} \times 0.05} = \mathbf{848\ V}$$

Current flowing through the material,

$$I = P/V \cos\phi = 400/848 \times 0.05 = \mathbf{9.48\ A}$$

Heat produced $\propto V^2 f$

$$\therefore \quad V_2^2 f_2 = V_1^2 f_1 \quad \text{or} \quad f_2 = f_1 (V_1/V_2)^2 = 40 \times 10^6 (848/700)^2 = \mathbf{58.7\ MHz}$$

Example 1.11 *. A plywood board of 0.5 × 0.25 × 0.02 metre is to be heated from 25 to 125°C in 10 minutes by dielectric heating employing a frequency of 30 MHz. Determine the power required in this heating process. Assume specific heat of wood 1500/J/kg/°C; weight of wood 600 kg/m^3 and efficiency of process 50%.*

Solution. Volume of plywood to be heated = 0.5 × 0.25 × 0.02 = 0.0025 m^3

Weight of plywood = 0.0025 × 600 = 1.5 kg

Heat required to raise the temperature of plywood board from 25° to 125°

= 1.5 × 1500 (125 – 25) = 2,25,000 J or W-s

or H = 2,25,000/60 × 60 = 62.5 Wh

Since heating is to be done in 10 minutes, power required = 62.5/(10/60) = 375 W

Since efficiency is 50%, power input = 375/0.5 = **700 W**

ELECTRIC WELDING

Definition of Welding

It is the process of joining two pieces of metal or non-metal at faces rendered plastic or liquid by the application of heat or pressure or both. Filler material may be used to effect the union.

Welding Processes

All welding processes fall into two distinct categories:

1. **Fusion Welding**: It involves melting of the parent metal. Examples are:

(i) Carbon arc welding, metal arc welding, electron beam welding, electroslag welding and electrogas welding which utilize electric energy and

(ii) Gas welding and thermit welding which utilize chemical energy for the melting purpose.

2. **Non-fusion Welding**: It does not involve melting of the parent metal. Examples are:

(i) Forge welding and gas non-fusion welding which use chemical energy.

(ii) Explosive welding, friction welding and ultrasonic welding etc., which use mechanical energy.

(iii) Resistance welding which uses electrical energy.

Proper selection of the welding process depends on the

a) kind of metals to be joined
b) cost involved
c) nature of products to be fabricated and
d) production techniques adopted.

The principal welding processes have been tabulated in figure 1.26.

Use of Electricity in Welding

Electricity is used in welding for generating heat at the point of welding in order to melt the material which will subsequently fuse and form the actual weld joint. There are many ways of producing this localized heat but the two most common methods are as follows:

1. **Resistance welding**—here current is passed through the inherent resistance of the joint to be welded thereby generating the heat as per the equation $I^2\ Rt/J$ kilocalories.

2. **Arc welding**—here electricity is conducted in the form of an arc which is established between the two metallic surfaces.

Formation and Characteristics of Electric Arc

An electric arc is formed whenever electric current is passed between two metallic electrodes which are separated by a short distance from each other. The arc is started by momentarily touching the positive electrode (anode) to the negative metal (or plate) and then withdrawing it to about 3 to 6 mm from the plate. When electrode first touches the plate, a large short-circuit current flows and as it is later withdrawn from the plate, current continues to flow in the form of a spark across the air gap so formed. Due to this spark (or discharge), the air in the gap becomes ionized *i.e.* is split into negative electrons and positive ions. Consequently, air becomes conducting and current is able to flow across the gap in the form of an arc.

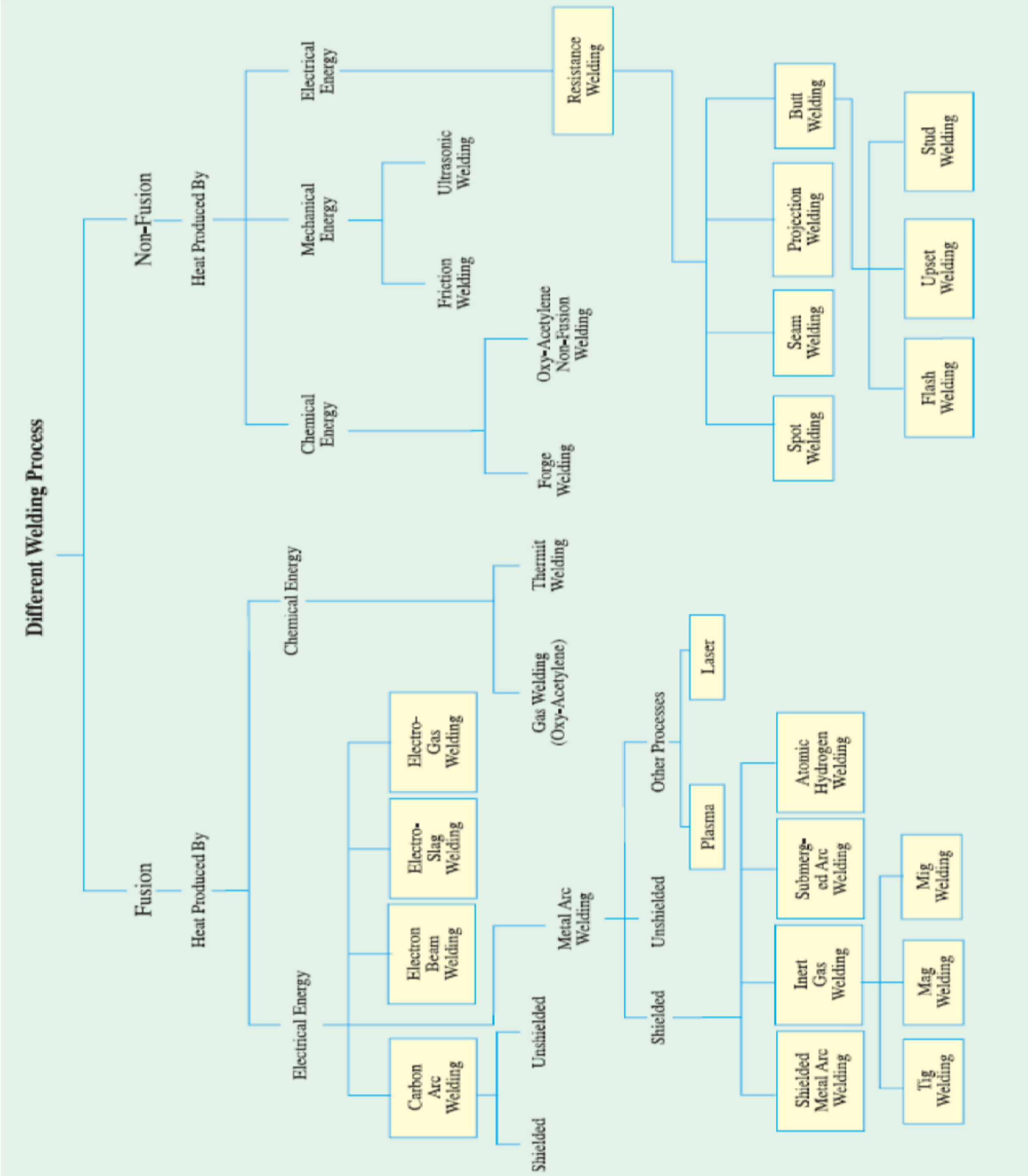


Fig. 1.26

As shown in figure 1.27, the arc consists of ***lighter*** electrons which flow from cathode to anode and ***heavier*** positive ions which flow from anode to cathode. Intense heat is generated when high velocity electrons strike the anode. Heat generated at the cathode is much less because of the low velocity of the impinging ions. It is found that nearly **two-third** of the heat is developed at the anode which burns into the form of a crater where temperature rises to a value of 3500-4000°C. The remaining one-third of the heat is developed near the cathode. The above statement is true in all d.c. systems of welding where positive side of the circuit is the hottest side. As a result, an electrode connected to the positive end of the d.c. supply circuit will burn 50% faster than if connected to the negative end. This fact can be used for obtaining desired penetration of the base metal during welding.

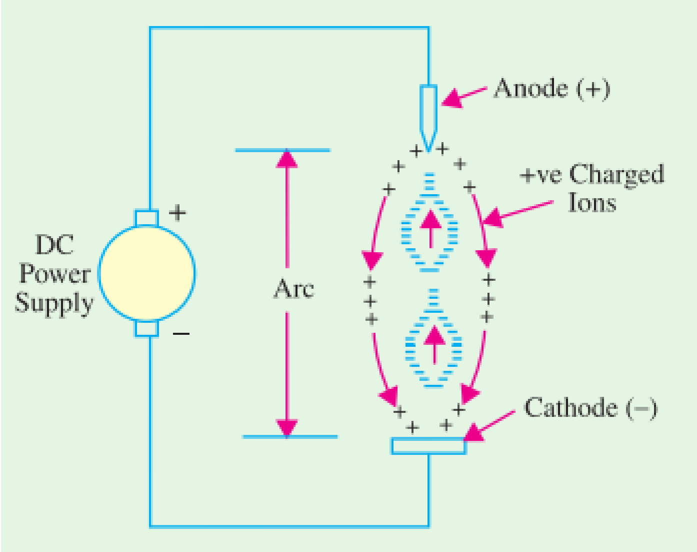


Fig. 1.26

If positive supply end is connected to the base metal (which is normally grounded), penetration will be greater due to more heat and, at the same time, the electrode will burn away slowly [Fig. 1.27(*a*)] since it is connected to the negative end of the supply. If supply connections are reversed, the penetration of heat zone in the base metal will be comparatively shallow and at the same time electrode will burn fast [Fig.1.27 (*b*)]. AC supply produces a penetration depth that is early halfway between that achieved by the d.c. positive ground and negative ground as shown in figure.1.27(*c*). It may be noted that with a.c. supply, heat is developed equally at the anode and cathode due to rapid reversal of their polarity. The arc utilized for arc welding is a low-voltage high-current discharge. The voltage required for striking the arc is higher than needed for maintaining it. Moreover, amperage increases as voltage decreases after the arc has been established. Figure 1.28 shows V/I characteristics of an electric arc for increasing air-gap lengths. The voltage required to strike a d.c. arc is about 50-55 V and that for a.c. arc is 80-90 V. The voltage drop across the arc is nearly 15-20 V. It is difficult to maintain the arc with a voltage less than 14 V or more than 40 V.

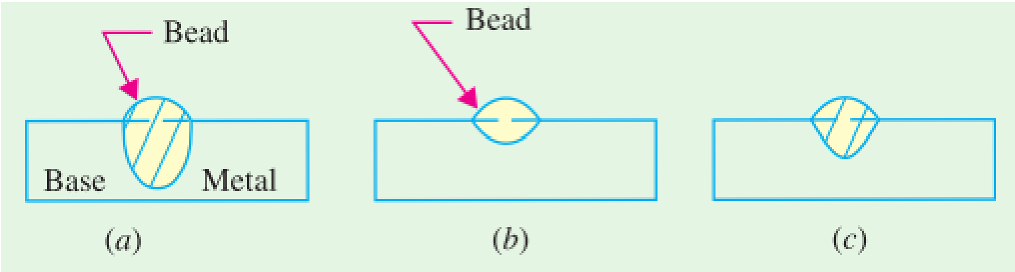


Fig. 1.27

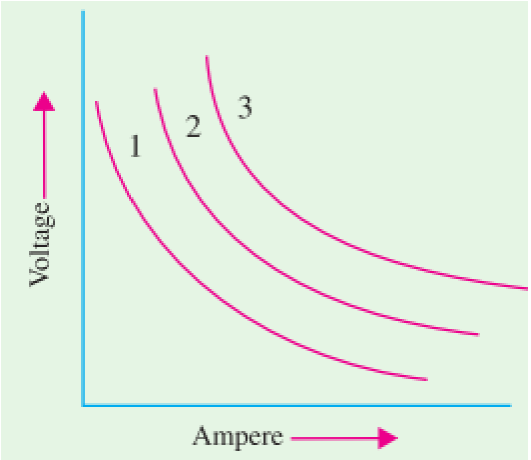


Fig. 1.28

Effect of Arc Length

In metal arc welding, a fairly short arc length is necessary for getting good welds. Short arc length permits the heat to be concentrated on the workpiece, is more stable because effect of magnetic blow is reduced and the vapours from the arc surround the electrode metal and the molten pool thereby preventing air from destroying the weld metal. When arc length is long

1. large amount of heat is lost into the surrounding area thus preventing good penetration and fusion;
2. arc flame is very unstable since effect of magnetic blow is increased. Hence, arc flame will have a tendency to blow out;
3. air is able to reach the molten globule of metal as it passes from the electrode to the weld and weld pool. It leads to the contamination of the weld due to absorption of oxygen and nitrogen;
4. weld deposits have low strength, poor ductility, high porosity, poor fusion and excessive spatter.

The length of arc required for welding will depend on the kind of electrode used, its coating, its diameter, position of welding and the amount of current used. Usually, shorter arc length are necessary for vertical, horizontal and overhead welding than for flat welding.

Arc Blow

An arc column can be considered as a flexible current-carrying conductor which can be easily deflected by the magnetic field set up in its neighbourhood by the positive and negative leads from the d.c. welding set. The two leads carry currents in the opposite directions and hence, set up a repulsive magnetic force which pulls the arc away from the weld point particularly when welding corners where field concentration is maximum. The deflection of the arc is called arc blow. This condition is encountered only with d.c. welding sets and is especially noticeable when welding with bare electrodes. It is experienced most when using currents above 200 A or below 40 A. Due to arc blow, heat penetration in the required area is low which leads to incomplete fusion and bead porosity apart from excessive weld spatter.

An electric arc is produced when electricity is passed between two electrodes

Arc blow can be avoided by using a.c. rather than d.c. welding machines because reversing currents in the welding leads produce magnetic fields which cancel each other out thereby eliminating the arc blow. However, with d.c. welding machines, arc blow effects can be minimized by

(i) welding away from the earth ground connection,
(ii) changing the position of the earth connection on the work,
(iii) wrapping the welding electrode cable a few turns around the work,
(iv) reducing the welding current or electrode size,
(v) reducing the rate of travel of the electrode and
(vi) shortening the arc column length etc.

Polarity in DC Welding

Arc welding with the electrode connected to the positive end of the d.c. supply is called reverse polarity.* Obviously, the workpiece is connected to the negative end.

A better name for d.c. reverse polarity (DCRP) is electrode-positive as shown in figure 1.29(a). As stated earlier in Art. 48.4, two-third of the arc heat is developed at the anode. Hence, in DCRP welding, electrode is the hottest whereas work piece is comparatively cooler. Consequently, electrode burns much faster but weld bead is relatively shallow and wide. That is why thick and heavily coated electrodes are used in DCRP welding because they require more heat for melting. Arc welding with the electrode connected to the negative end of the d.c. supply is called straight polarity.** Obviously, the workpiece is connected to the positive end as shown in figure 1.29 (b). A better name for d.c. straight polarity (DCSP) is electrode-negative.

* In British literature, it is called straight polarity.
** In British literature, it is called reverse polarity.

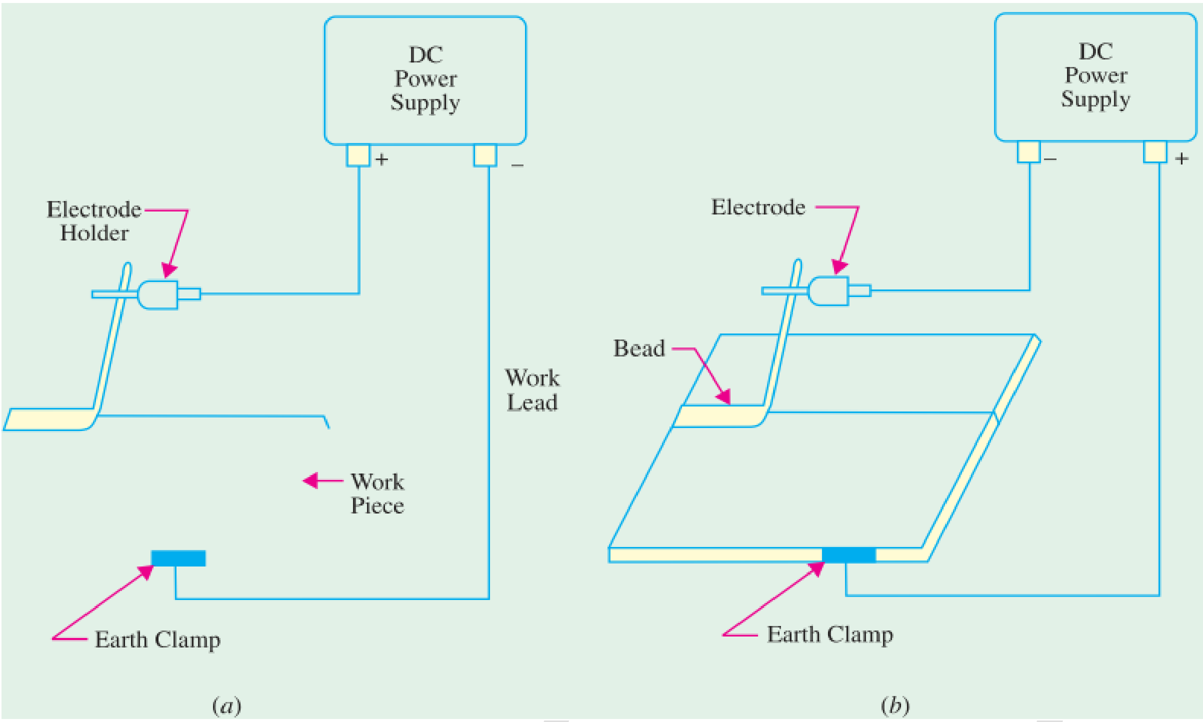


Fig. 1.29

In DCSP welding, workpiece is the hottest, hence base metal penetration is narrow and deep. Moreover, bare and medium-coated electrodes can be used in this welding as they require less amount of heat for melting. It is seen from the above discussion that polarity necessary for the welding operation is determined by the type of electrode used. It is also worth noting that in a.c. welding, there is no choice of polarity because the circuit becomes alternately positive, first on one side and then on the other. In fact, it is a combination of DCSP and D CRP.

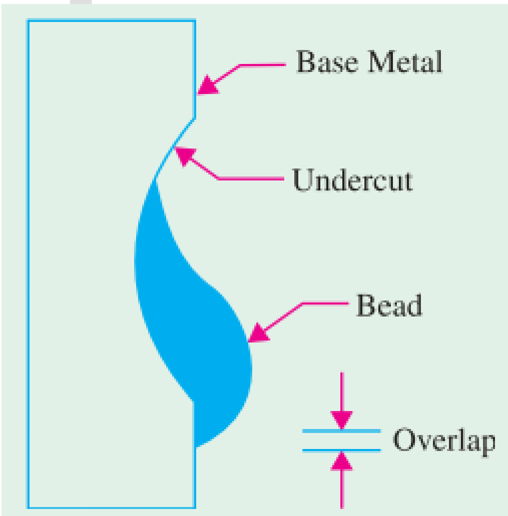


Fig. 1.30

Four Positions of Arc Welding

There are four basic positions in which manual arc welding is done.

1. **Flat position.** It is shown in figure 1.31 (a). Of all the positions, flat position is the easiest, most economical and the most used for all shielded arc welding. It provides the strongest

weld joints. Weld beads are exceedingly smooth and free of slag spots. This position is most adaptable for welding of both ferrous and non-ferrous metals particularly for cast iron.

2. **Horizontal Position**. It is the second most popular position and is shown in figure 1.31(b). It also requires a short arc length because it helps in preventing the molten puddle of the metal from sagging. However, major errors that occur while welding in horizontal position are under- cutting and over-lapping of the weld zone (Fig. 1.30).

3. **Vertical Position**. It is shown in figure 1.31 (c). In this case, the welder can deposit the bead either in the uphill or downhill direction. Downhill welding is preferred for thin metals because it is faster than the uphill welding. Uphill welding is suited for thick metals because it produces stronger welds.

4. **Overhead Position**. It is shown in figure 1.31 (d). Here, the welder has to be very cautious otherwise he may get burnt by drops of falling metal. This position is thought to be the most hazardous but not the most difficult one.

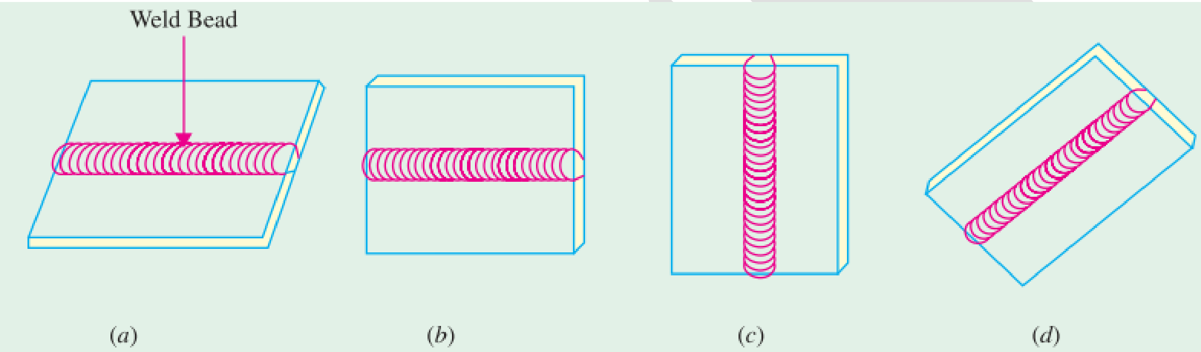


Fig. 1.31

Electrodes for Metal Arc Welding

An electrode is a filler metal in the form of a wire or rod which is either bare or coated uniformly with flux. As per IS : 814-1970, the contact end of the electrode is left bare and clean to a length of 20-30 mm. for inserting it into electrode holder (Fig. 1.32).

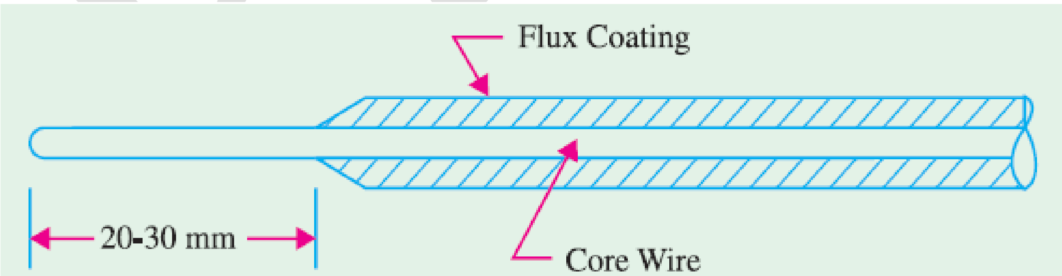


Fig. 1.32

Metal arc welding was originally done with bare electrodes which consisted of a piece of wire or rod of the same metal as the base metal. However, due to atmospheric contamination, they produced brittle and poor quality welds. Hence, bare wire is no longer used except for automatic welding in which case arrangement is made to protect the weld area from the atmosphere by either powdered flux or an inert gas. Since 1929, coated electrodes are being extensively used for shielded arc welding.

They consist of a metal core wire surrounded by a thick flux coating applied by extrusion, winding or other processes. Depending on the thickness of the flux coating, coated electrodes may be classified into ***(i)*** lightly-dusted (or dipped) electrodes and ***(ii)*** semi-coated (or heavycoated) electrodes. Materials commonly used for coating are ***(i)*** titanium oxide ***(ii)*** ferromanganese ***(iii)*** silica flour ***(iv)*** asbestos clay ***(v)*** calcium carbonate and ***(vi)*** cellulose with sodium silicate often used to hold ingredients together.

Electrode coating contributes a lot towards improving the quality of the weld. Part of the coating burns in the intense heat of the arc and provides a gaseous shield around the arc which prevents oxygen, nitrogen and other impurities in the atmosphere from combining with the molten metal to cause a poor quality brittle and weak weld. Another portion of the coating flux melts and mixes with the impurities in the molten pool causing them to float to the top of the weld where they cool in the form of slag (Fig. 1.33). This slag improves the bead quality by protecting it from the contaminating effects of the atmosphere and causing it to cool down more uniformly. It also helps in controlling the basic shape of the weld bead.

The type of electrode used depends on the type of metal to be welded, the welding position, the type of electric supply whether a.c. or d.c. and the polarity of the welding machine.

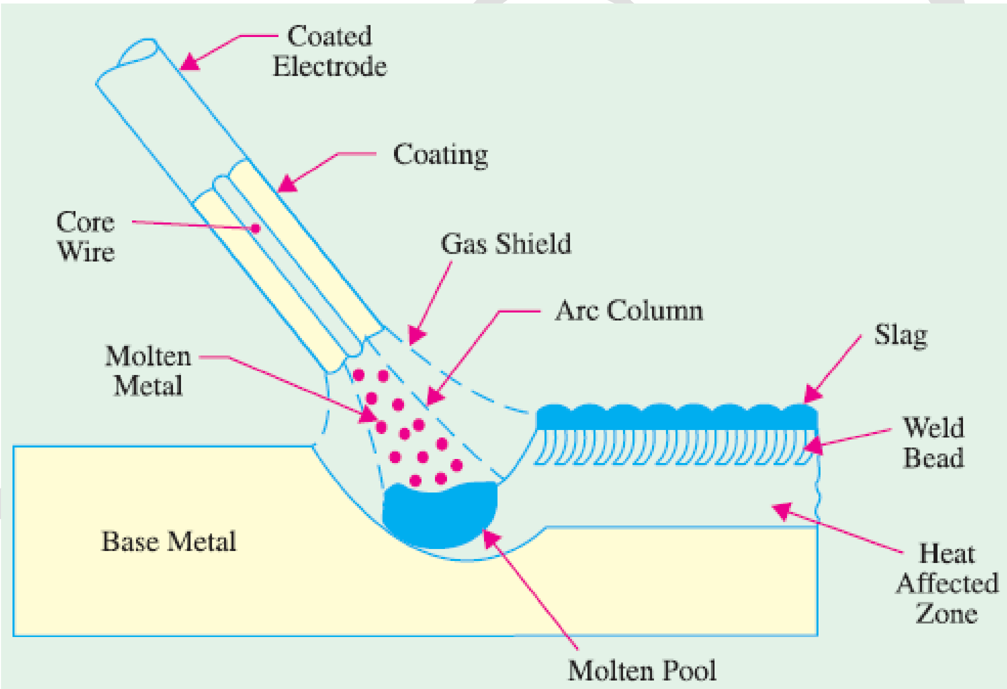


Fig. 1.33

Advantages of Coated Electrodes

The principal advantages of using electrode coating are as under:

1. It stabilizes the arc because it contains ionizing agents such as compounds of sodium and potassium.
2. It fluxes away impurities present on the surface being welded.
3. It forms slag over the weld which ***(i)*** protects it from atmospheric contamination
   (ii) makes it cool uniformly thereby reducing the changes of brittleness and
   (iii) provides a smoother surface by reducing _ripples‘ caused by the welding operation.
4. It adds certain materials to the weld metal to compensate for the loss of any volatile alloying

elements or constituents lost by oxidization.

5. It speeds up the welding operation by increasing the rate of melting.
6. It prevents the sputtering of metal during welding.
7. it makes it possible for the electrode to be used on a.c. supply. In a.c. welding, arc tends to cool and interrupt at zero-current positions. But the shielding gases produced by the flux keep the arc space ionized thus enabling the coated electrodes to be used on a.c. supply. It is worth noting that efficiency of all coated (or covered) electrodes is impaired by dampness. Hence, they must always be stored in a dry space. If dampness is suspected, the electrodes should be dried in a warm cabinet for a few hours.

Types of Joints and Types of Applicable Welds

Bureau of Indian Standards (B.I.S.) has recommended the following types of joints and the welds applicable to each one of them (Fig. 1.34).

1. Tee joint — with six types of welds.
2. Corner joint — with two types of welds.
3. Edge joint — with one type of weld.
4. Lap joint — with four types of welds.
5. Butt joint — with nine types of welds.

Arc Welding Machines

Welding is never done directly from the supply mains. Instead, special welding machines are used which provided currents of various characteristics. Use of such machines is essential for the following reasons :

1. To convert a.c. supply into d.c. supply when d.c. welding is desired.
2. To reduce the high supply voltage to a safer and suitable voltage for welding purposes
3. To provide high current necessary for arc welding without drawing a corresponding high current from the supply mains.
4. To provide suitable voltage/current relationships necessary for arc welding at minimum cost.

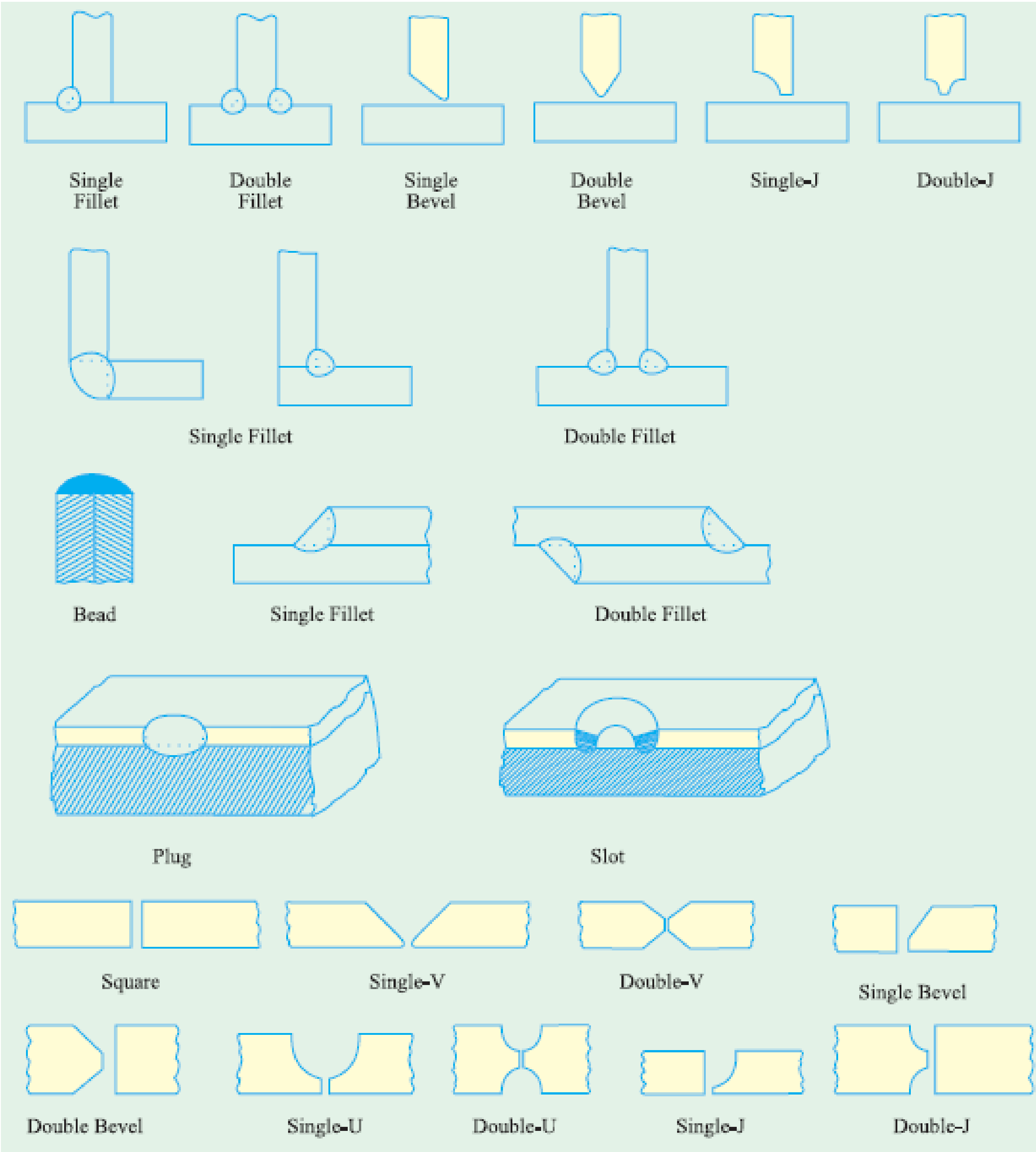


Fig. 1.34

There are two general types of arc welding machines :

(*a*) d.c. welding machines

(i) motor-generator set

(ii) a.c. transformers with rectifiers

(*b*) a.c. welding machines

V-I Characteristics of Arc Welding DC Machines

It is found that during welding operation, large fluctuations in current and arc voltage result from the mechanism of metal transfer and other factors. The welding machine must compensate for such changes in arc voltage in order to maintain an even arc column. There are three major voltage/current characteristics used in modern d.c. welding machines which help in controlling these current fluctuations:

1. drooping arc voltage (DAV).
2. constant arc voltage (CAV).
3. rising arc voltage (RAV).

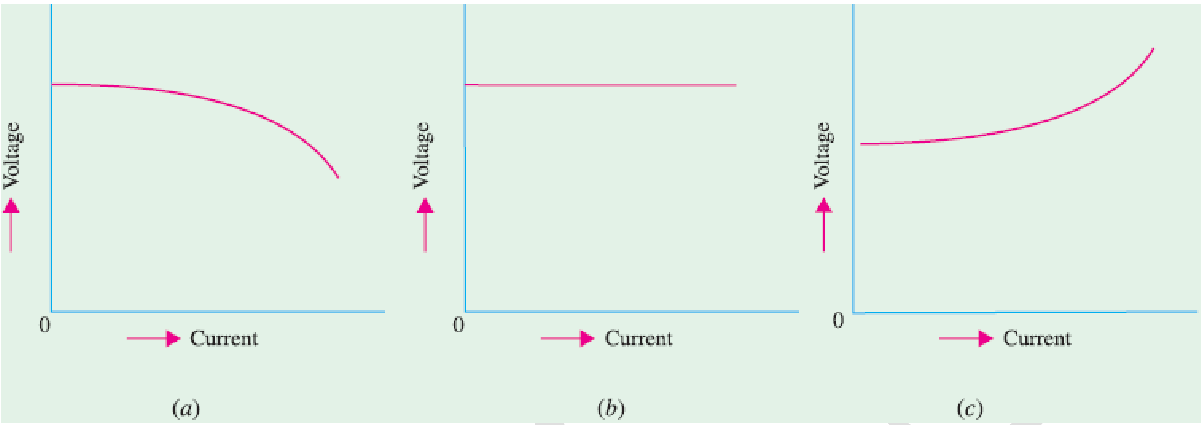


Fig. 1.35

The machines with DAV characteristics have high open-circuit voltage which drops to a minimum when arc column is started. The value of current rises rapidly as shown in figure 1.35(*a*). This type of characteristic is preferred for manual shield metal arc welding.

The CAV characteristic shown in Fig. 1.35 (b) is suitable for semi-automatic or automatic welding processes because voltage remains constant irrespective of the amount of current drawn. Because of its rising voltage characteristic, RAV has an advantage over CAV because it maintains a constant arc gap even if short circuit occurs due to metal transfer by the arc. Moreover, it is well adopted to fully automatic process. DC welding machines can be controlled by a simple rheostat in the exciter circuit or by a combination of exciter regulator and series of field taps. Some arc welding are equipped with remote-controlled current units enabling the operator to vary voltage amperage requirement without leaving the machines.

DC Welding Machines with Motor Generator Set

Such a welding plant is a self-contained single-operator motor-generator set consisting of a reverse series winding d.c. generator driven by either a d.c. or an a.c. motor (usually 3-phase). The series winding produces a magnetic field which opposes that of the shunt winding. On open-circuit, only shunt field is operative and provides maximum voltage for striking the arc. After the arc has been established, current flows through the series winding and sets up a flux which opposes the flux produced by shunt winding. Due to decreases in the net flux, generator voltage is decreased (Art. 1.33). With the help of shunt regulator, generator voltage and current values can be adjusted to the desired level. Matters are so arranged that despite changes in arc voltage due to variations in arc length, current remains practically constant. Figure 1.36 shows the circuit of a d.c. motor-generator type of welding machine.

Advantages: Such a d.c. welder has the following advantages:

1. It permits portable operation.
2. It can be used with either straight or reverse polarity.
3. It can be employed on nearly all ferrous and non-ferrous materials.
4. It can use a large variety of stick electrodes.
5. It can be used for all positions of welding.

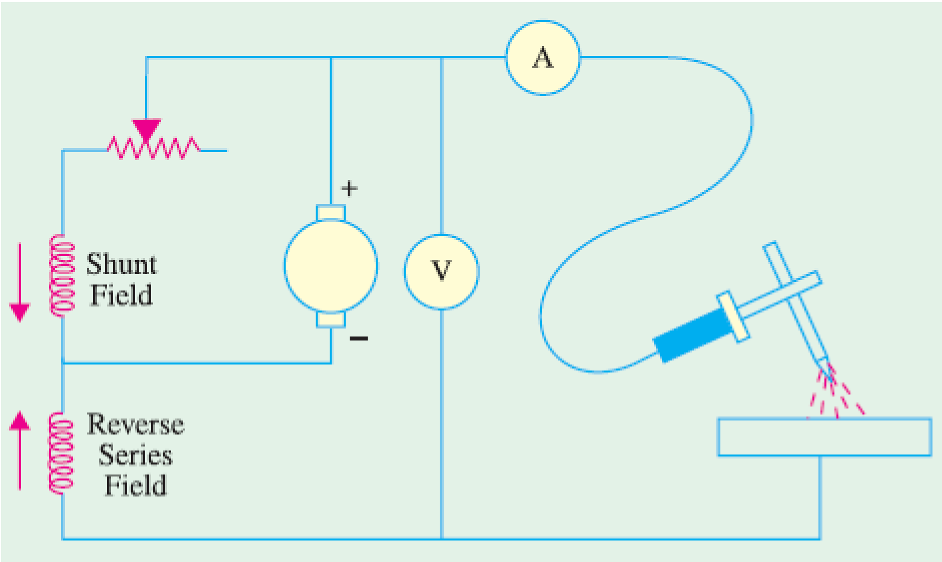


Fig. 1.36

Disadvantages:

1. It has high initial cost.
2. Its maintenance cost is higher.
3. Machine is quite noisy in operation.
4. It suffers from arc blow.

AC Rectified Welding Unit

It consists of a transformer (single-or three-phase) and a rectifier unit as shown in figure 1.37. Such a unit has no moving parts, hence it has long life. The only moving part is the fan for cooling the transformer. But this fan is not the basic part of the electrical system. Figure 1.37 shows a single-phase full-wave rectified circuit of the welder. Silicon diodes are used for converting a.c. into d.c. These diodes are hermetically sealed and are almost ageless because they maintain rectifying characteristics indefinitely. Such a transformer-rectifier welder is most adaptable for shield arc welding because it provides both d.c. and a.c. polarities. It is very efficient and quiet in operation. These welders are particularly suitable for the welding of ***(i)*** pipes in all positions ***(ii)*** non-ferrous metals ***(iii)*** low-alloy and corrosion- heat and creep-resisting steel ***(iv)*** mild steels in thin gauges.

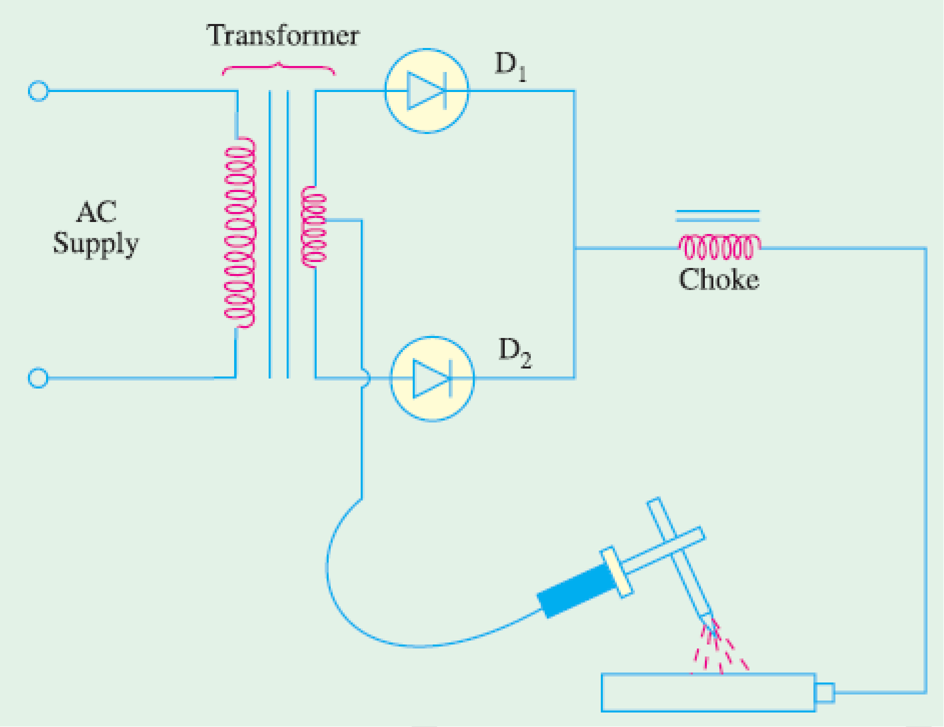


Fig. 1.37

AC Welding Machines

As shown in figure 1.38, it consists of a step-down transformer with a tapped secondary having an adjustable reactor in series with it for obtaining drooping V/I characteristics. The secondary is tapped to give different voltage/ current settings.

Advantages: This a.c. welder which can be operated from either a single-phase or 3-phase supply has the following advantages: ***(i)*** Low initial cost ***(ii)*** Low operation and maintenance cost ***(iii)*** Low wear ***(iv)*** No arc blow

Disadvantages. (*i*) its polarity cannot be changed **(*ii*)** it is not suitable for welding of cast iron and non-ferrous metals.

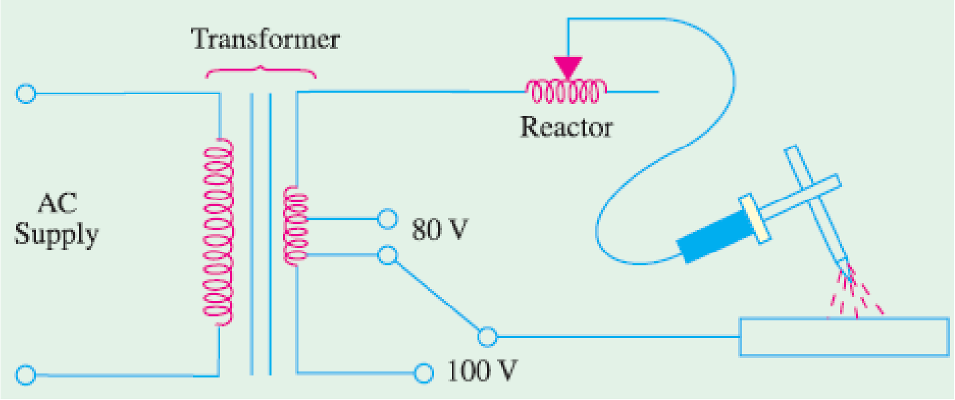


Fig. 1.38

Duty Cycle of a Welder

The duty cycle of an arc welder is based on a working period of 10 minutes. For example,if a welder is operated for 2 minutes in a period of 10 minutes, then its percentage duty cycle is (2/10) × 100 = 20 percent. Conversely, a 10 percent duty cycle would mean that the welder would be operated for 10 percent of 10 minutes i.e. for one minute only in a period of 10 minutes.

Usually, values of maximum amperage and voltage are indicated along with the duty cycle. It is advisable to adhere to these values. Suppose a welding machine has maximum amperage of 300A and voltage of 50 V for a duty cycle of 60 percent. If this machine is operated at higher settings and for periods longer than 6 minutes, then its internal insulation will deteriorate and cause its early failure.

Carbon Arc Welding

(a) General

Carbon arc welding was the first electric welding process developed by a French inventor Auguste de Meritens in 1881. In this process, fusion of metal is accomplished by the heat of an electric arc. No pressure is used and generally, no shielding atmosphere is utilized. Filler rod is used only when necessary. Although not used extensively these days, it has, nevertheless, certain useful fields of application.

Carbon arc welding differs from the more common shield metal arc welding in that ***it uses nonconsumable carbon or graphic electrodes*** instead of the consumable flux-coated electrodes.

(b) Welding Circuit

The basic circuit is shown in figure 1.39 and can be used with d.c. as well as a.c. supply. When direct current is used, the electrode is mostly negative (DCSP). The process is started by adjusting the amperage on the d.c. welder, turning welder ON and bringing the electrode into contact with the workpiece. After the arc column starts, electrode is withdrawn 25 – 40 mm away and the arc is maintained at this distance. The arc can be extinguished by simply removing the electrode from the workpiece completely. The only function of the carbon arc is to supply heat to the base metal. This heat is used to melt the base metal or filler rod for obtaining fusion weld Depending on the type and size of electrodes, maximum current values range from 15 A to 600 A for single-electrode carbon arc welding.

(c) Electrodes

These are made of either carbon or graphite, are usually 300 mm long and 2.5 – 12mm in diameter. Graphite electrodes are harder, more brittle and last longer than carbon electrodes. They can withstand higher current densities but their arc column is harder to control. Though considered nonconsumable, they do disintegrate gradually due to vaporisation and oxidisation.

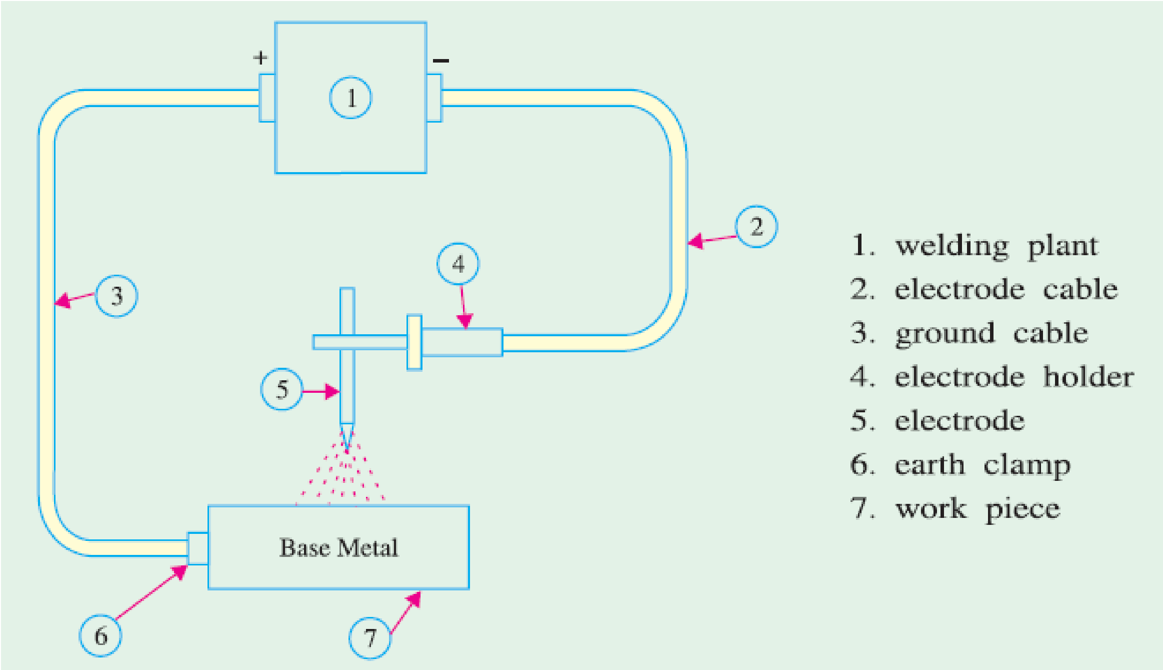


Fig. 1.39

(*d*) Applications

1. The joint designs that can be used with carbon arc welding are butt joints, bevel joints, flange joints, lap joints and fillet joints.
2. This process is easily adaptable for automation particularly where amount of weld deposit is large and materials to be fabricated are of simple geometrical shapes such as water tanks.
3. It is suitable for welding galvanised sheets using copper-silicon-manganese alloy filler metal.
4. It is useful for welding thin high-nickel alloys.
5. Monel metal can be easily welded with this process by using a suitable coated filler rod.
6. Stainless steel of thinner gauges is often welded by the carbon-arc process with excellent results.

(*e*) Advantages and Disadvantages

5. The main advantage of this process is that the temperature of the molten pool can be easily controlled by simply varying the arc length.
6. It is easily adaptable to automation.
7. It can be easily adapted to inert gas shielding of the weld and
8. It can be used as an excellent heat source for brazing, braze welding and soldering etc.

Its disadvantages are as under:

1. A separate filler rod has to be used if any filler material is required.
2. Since arc serves only as a heat source, it does not transfer any metal to help reinforce the weld joint.
3. The major disadvantage of the carbon-arc process is that blow holes occur due to magnetic arc blow especially when welding near edges of the workpiece.

Submerged Arc Welding

In this ***fusion*** process, welding is done under a blanket of granulated flux which shields the weld from all bad effects of atmospheric gases while a consumable electrode is continuously and mechanically fed into the arc. The arc, the end of the bare metal electrode and the molten weld pool are all submerged under a thick mound of finely-divided granulated powder that contains deoxidisers, cleansers and other fluxing agents. The fluxing powder is fed from a hopper that is carried on the welding head itself (Fig. 1.40). This hopper spread the powder in a continuous mound ahead of the electrode in the direction of welding. Since arc column is completely submerged under the powder, there is no splatter or smoke and, at the same time, weld is completely protected from atmospheric contamination. Because of this protection, weld beads are extremely smooth. The flux adjacent to the arc column melts and floats to the top of the molten pool where it solidifies to form slag. This slag is easy to remove. Often it cracks off by itself as it cools. The unused flux is removed and is reused again and again.

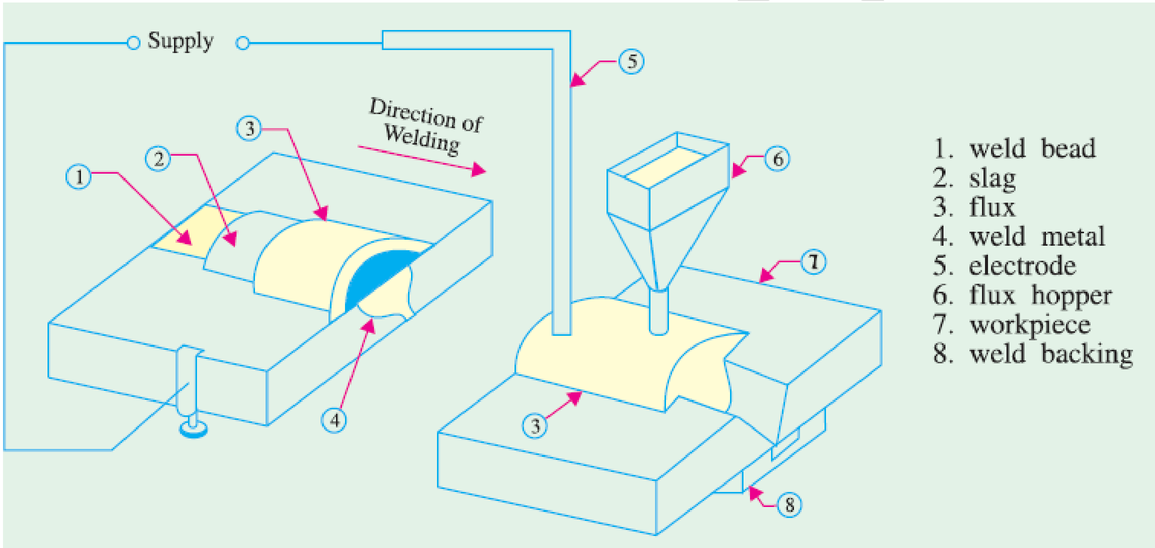


Fig. 1.40

The electrode is either a bare wire or has a slight mist of copper coated over it to prevent oxidation. In automatic or semi-automatic submerged arc welding, wire electrode is fed mechanically through an electrically contacting collet. Though a.c. power supply may be used, yet d.c. supply is more popular because it assures a simplified and positive control of the welding process.

This process requires high current densities about 5 to 6 times of those used in ordinary manual stick electrode welding. As a result, melting rate of the electrode as well as welding speed becomes much higher. Faster welding speed minimizes distortion and warpage.

The submerged arc process is suitable for

1. Welding low-alloy, high-tensile steels.
2. Welding mild, low-carbon steels.
3. Joining medium-carbon steel, heat-resistant steels and corrosion-resistant steels etc.
4. Welding nickel, Monel and other non-ferrous metals like copper.

This process has many industrial applications such as fabrication of pipes, boiler pressure vessels, railroad tank cars, structural shapes etc. which demand welding in a straight line. Welds made by this process have high strength and ductility. A major advantage of this process is that fairly thick sections can be welded in a single pass without edge preparation.

Submerged arc welding can be done manually where automatic process is not possible such as on curved lines and irregular joints. Such a welding gun is shown in figure 1.41. Both manual and automatic submerged arc processes are most suited for flat and slightly downhill welding positions.

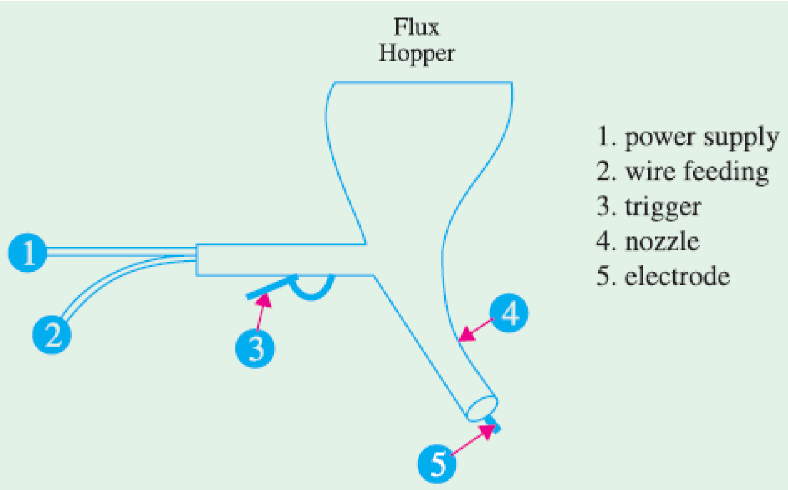


Fig. 1.41

Twin Submerged Arc Welding

As shown in figure 1.42 in this case, two electrodes are used simultaneously instead of one. Hence, weld deposit size is increased considerably. Moreover, due to increase in welding current (upto 1500 A), much deeper penetration of base metal is achieved.

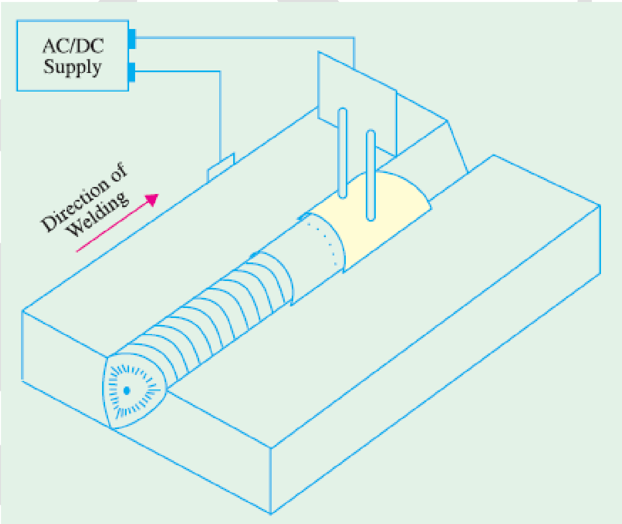


Fig. 1.41

Gas Shield Arc Welding

In this fusion process, welding is done with bare electrodes but weld zone is shielded from the atmosphere by a gas which is piped to the arc column. Shielding gases used are carbon dioxide, argon, helium, hydrogen and oxygen. No flux is required. Different processes using shielding gas are as follows.

(a) Tungsten inert-gas (TIG) Process

In this process, non-consumable tungsten electrode is used and filler wire is fed separately. The weld zone is shielded from the atmosphere by the inert gas (argon or helium) which is ducted directly to the weld zone where it surrounds the tungsten and the arc column.

(b) Metal inert-gas (MIG) Process

It is a refinement of the TIG process. It uses a bare consumable (*i.e.* fusible) wire electrode which acts as the source for the arc column as well as the supply for the filler material. The weld zone is shielded by argon gas which is ducted directly to the electrode point.

TIG Welding

(a) Basic Principle

It is an electric process which uses a bare non-consumable tungsten electrode for striking the arc only (Fig. 1.42). Filler material is added separately. It uses an inert gas to shield the weld puddle from atmospheric contamination. This gas is ducted directly to the weld zone from a gas cylinder.

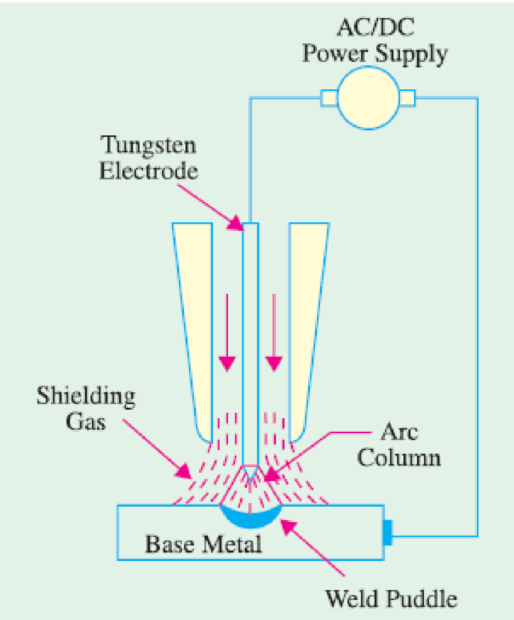


Fig. 1.42

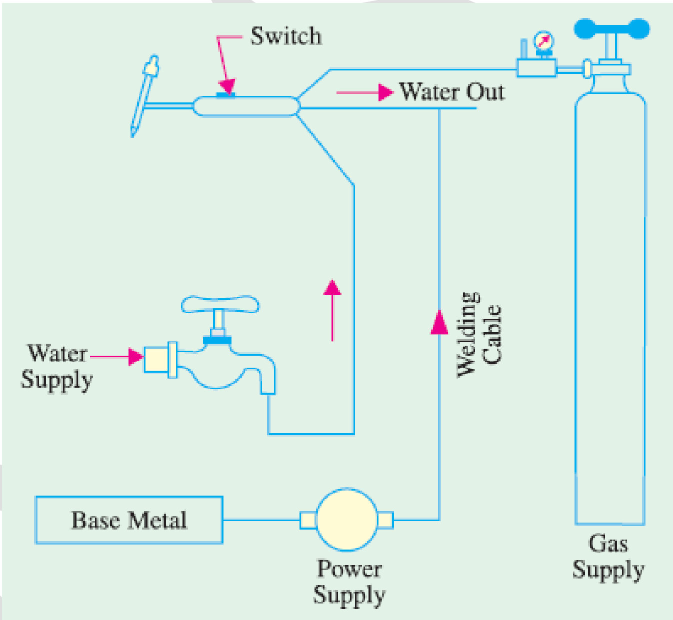


Fig. 1.43

(b) Welding Equipment

The usual TIG welding system consists of the following (Fig. 1.43).

1. A standard shield arc welding machine complete with cables etc.
2. A supply of inert gas complete with hose, regulators etc.
3. A source of water supply (in the case of water-cooled torches).
4. A TIG torch with a control switch to which all the above are connected.

(c) Electrodes

The electrodes are made of either pure tungsten or zirconiated or thoriated tungsten. Addition of zironium or thorium (0.001 to 2%) improves electron emission tremendously.

(d) Power Supply

The three basic power supplies used in TIG operation are :

1. DCSP power supply–here electrode is negative, runs cooler and hence can be thin.
2. DCRP power supply–here electrode is positive and hot. Hence, it has to be large.

3. A.C. high frequency (ACHF) power supply–it is a combination of standard a.c. supply of 50 Hz and high-voltage high-frequency d.c. supply. The function of this d.c. supply is to sustain the arc when a.c. supply is at zero current positions.

(*e*) Advantages of TIG Welding

1. It provides maximum protection to weld bead from atmospheric contamination.
2. TIG welds are stronger, more ductile and more corrosion-resistant than those of shield metal arc welding.
3. Since no flux is used, there is no flux entrapment in the bead.
4. Since no flux is required, a wider variety of joint designs can be used.
5. No post-weld cleansing is necessary.
6. There is no weld splatter or sparks that could damage the surface of the base metal.
7. It gives relatively fast welding speeds.
8. It is suitable for welding food or medical containers where entrapment of any decaying organic matter could be extremely harmful.
9. It is suitable for all welding positions–the flat, horizontal, vertical and overhead positions.

The joints suitable for TIG welding process are **(*i*)** butt joint **(*ii*)** lap joint **(*iii*)** T-joint, **(*i*v**) corner joint and **(*v*)** edge joint.

(*f*) Applications

1. Aluminum and its alloys — AC/DCRP
2. Magnesium and its alloys — ACHF
3. Stainless steel — DCSP
4. Mild steel, low-alloy steel, medium-carbon steel and cast iron — DCSP
5. Copper and allows — DCSP
6. Nickel and alloys — DCSP

TIG welding is also used for dissimilar metals, hard facing and surfacing of metals. Special industrial applications include manufacture of metal furniture and air-conditioning equipment.

Figure 1.44 shows Phillips 400-D compact fan-cooled DC TIG welding set which has an open-circuit voltage of 80 V and a welding current of 400 A with 60% duty cycle and 310 A with 100% duty cycle.

Fig. 1.44 (TIG welding set)

MIG Welding

(a) Basic Principle

It is also called inert-gas consumable- electrode process. The fusible wire electrode is driven by the drive wheels. Its function is two-fold: to produce arc column and to provide filler material. This process uses inert gas for shielding the weld zone from atmospheric contamination. Argon is used to weld non-ferrous metals though helium gives better control of porosity and arc stability. This process can deposit large quantities of weld metal at a fast welding speed. The process is easily adaptable to semi-automatic or fully automatic operations.

(b) Welding Equipment

The basic MIG welding system (Fig. 1.45) consists of the following :

1. Welding power supply
2. Inert gas supply with a regulator and flow meter
3. Wire feed unit containing controls for wire feed, gas flow and the ON/OFF switch for MIG torch
4. MIG torch
5. Depending on amperage, a water cooling unit.

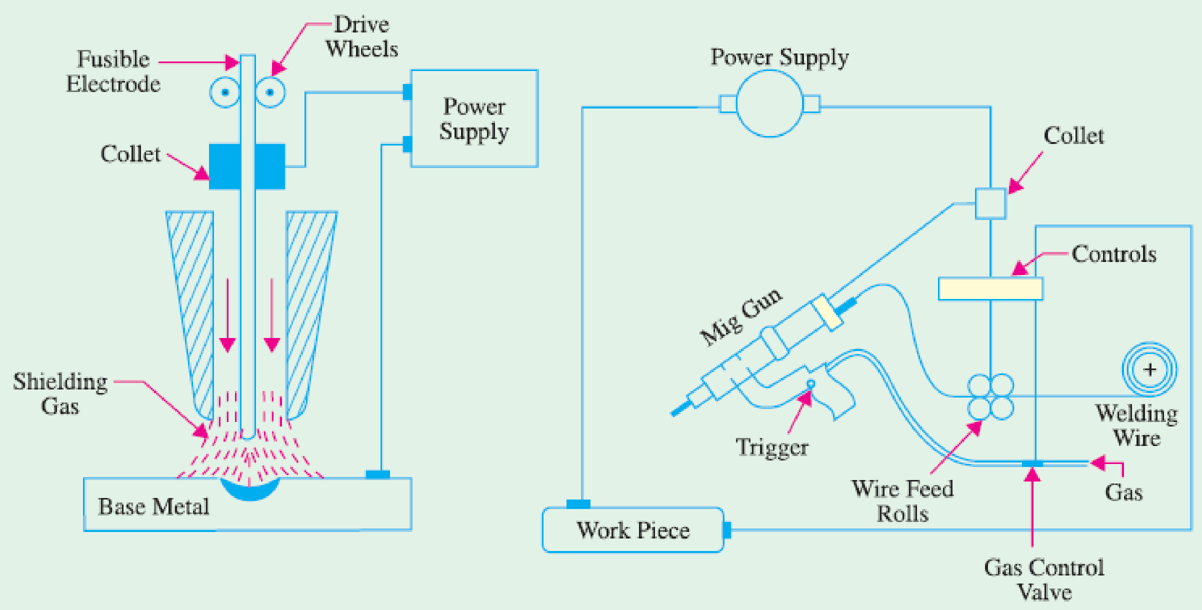


Fig. 1.45

(c) Electrode

It is a bare wire fed to the MIG gun by a suitable wire-feed mechanism.

(d) Power Supply

The major power supply used for MIG welding is DCRP and the machines which provide this supply are motor-generator sets or a.c. transformers with rectifiers. They have either CAV or RAV characteristics. The CAV supply gives the operator great latitude in arc length and is helpful in preventing the wire electrode from stubbing. A DCRP current produces deeper penetration and a cleaner weld surface than other types of current. The RAV machines are more suitable for automatic operation. They are capable of handling large diameter wires than CAV machines. Figure 1.46 shows semi-automatic forced-air cooled arc welding set MIG-400.

Fig. 1.46 MIG-400 Welding Set

It consists of

(i) Indarc 400 MMR rectifier which is basically a 3-phase transformer rectifier with silicon diodes and a constant potential output. It provides maximum current of 400 A at 40 V for 75% duty cycle and 350 A at 42 V for 100% duty cycle.

(ii) Indarc Wire Feeder which has a twin roll drive system, designed to feed 0.8 to 2.4 mm diameter welding wires to a hand-operated MIG welding torch.

(iii) MIG Torches which are available in both air-cooled and water-cooled varieties. Figure 1.47(*a*) and (*b*) show light-weight swan-necked torches which are designed to operate upto 360 A and 400 A with CO_2 as shielding gas. Figure 1.47 (*c*) shows a heavy-duty water-cooled torch designed to operate upto 550 A with CO_2 /mixed shielding gases at 100% duty cycle.

(iv) CO_2 Kit for hard wire applications and Argon Kit for soft wire applications.

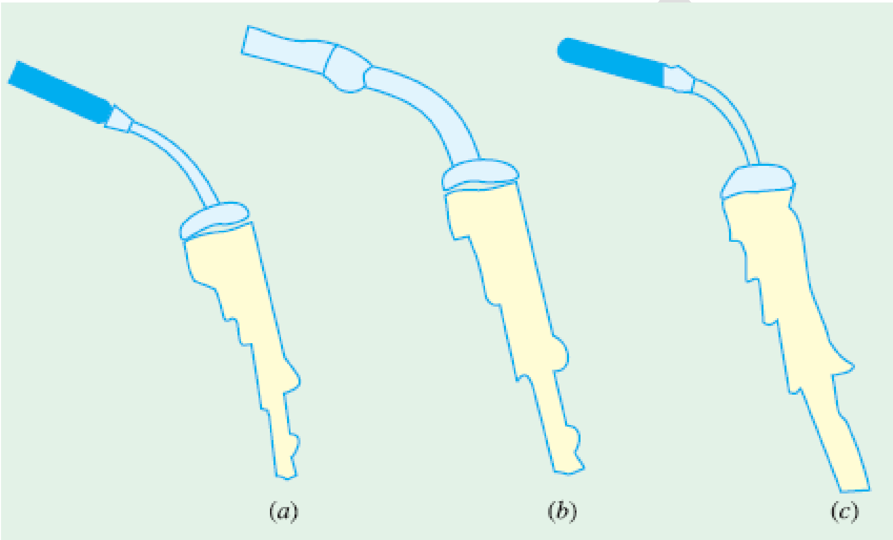


Fig. 1.47

(*e*) Advantages of MIG Welding

1. Gives high metal deposit rates varying from 2 to 8 kg/h.
2. Requires no flux.
3. Requires no post-welding cleaning.
4. Gives complete protection to weld bead from atmospheric contamination.
5. Is adaptable for manual and automatic operations.
6. Can be used for a wide range of metals both ferrous and non-ferrous.
7. Is easy to operate requiring comparatively much less operating skill.
8. Is especially suited for horizontal, vertical and overhead welding positions.

(*f*) Applications

With inert gas shielding, this process is suitable for fusion welding of **(*i*)** aluminium and its alloys **(*ii*)** nickel and its alloys **(*iii*)** copper alloys **(*iv*)** carbon steels **(*v*)** low-alloy steels **(*vi*)** high strength steels and **(*vii*)** titanium.

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Atomic Hydrogen Welding

(a) General

It is a non-pressure fusion welding process and the welder set is used only as heat supply for the base metal. If additional metal is required, a filler rod can be melted into the joint. It uses two tungsten electrodes between which an arc column (actually, an arc fan) is maintained by an a.c. supply.

(b) Basic Principle

As shown in figure 1.48, an arc column is struck between two tungsten electrodes with an a.c. power supply. Soon after, normal molecular hydrogen (H_2) is forced through this arc column. Due to intense heat of the arc column, this diatomic hydrogen is dissociated into atomic hydrogen (H).

However, atomic hydrogen being unstable recombines to form stable molecular hydrogen. In so doing, it releases intense heat at about 3750°C which is used to fuse the metals.

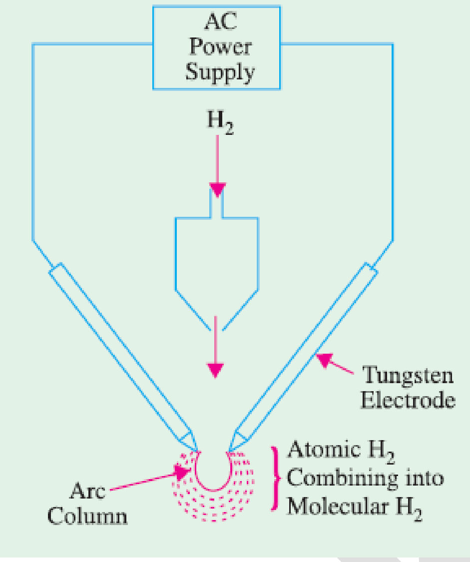


Fig. 1.47

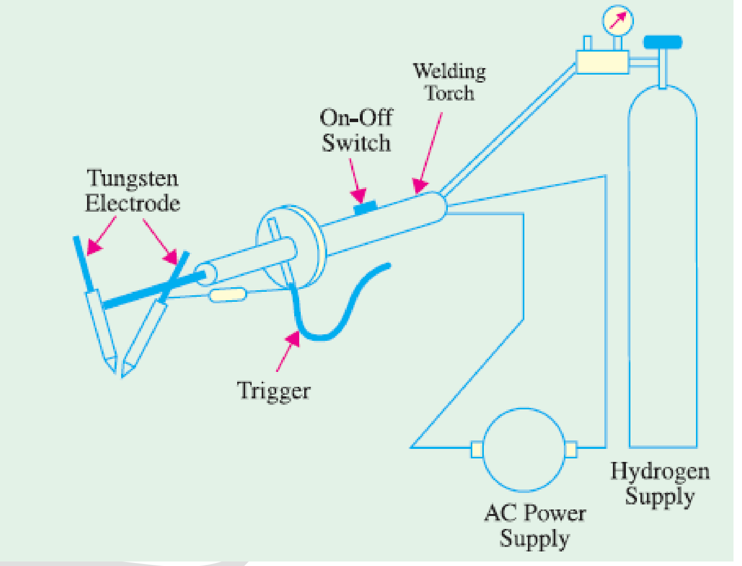


Fig. 1.48

(c) Welding Equipment

The welding equipment essentially consists of the following:

1. Standard welding machine consisting of a step-down transformer with tapped secondary (not shown in Fig. 1.48) energised from normal a.c. supply. Amperage requirement ranges from 15 A to 150 A
2. Hydrogen gas supply with an appropriate regulator
3. Atomic hydrogen welding torch having an ON-OFF switch and a trigger for moving the two tungsten electrodes close together for striking and maintaining the arc column

(d) Method of Welding

The torch is held in the right hand with first finger resting lightly on the trigger. The arc is struck either by allowing the two tungsten electrodes to touch and separate or by drawing the separated electrodes over a carbon block. At the same time, a stream of hydrogen is allowed to flow through the arc. As soon as the arc strikes, an intensely hot flame extends fanwise between the electrodes.

When this fan touches the workpiece, it melts it down quickly. If filler material is required, it can be added from the rod held in the left hand as in gas welding.

(e) Advantages

1. Arc and weld zone are shrouded by burning hydrogen which, being an active reducing agent, protects them from atmospheric contamination.
2. Can be used for materials too thin for gas welding.
3. Can weld quite thick sections.
4. Gives strong, ductile and sound welds.
5. Can be used for welding of mild steel, alloy steels and stainless steels and aluminum alloys.
6. Can also be used for welding of most non-ferrous metals such as nickel, monel, brass, bronze, tungsten and molybdenum etc.

Resistance Welding

It is fundamentally a heat and squeeze process. The term **'*resistance welding*'** denotes a group of processes in which welding heat is produced by the resistance offered to the passage of electric current through the two metal pieces being welded. These processes differ from the fusion processes in the sense that no extra metal is added to the joint by means of a filler wire or electrode. According to Joule's law, heat produced electrically is given by

$H = I^2Rt$ J.

Obviously, amount of heat produced depends on.

(*i*) square of the current **(*ii*)** the time of current and **(*iii*)** the resistance offered.

As seen, in simple resistance welding, high-amperage current is necessary for adequate weld. Usually, R is the contact resistance between the two metals being welded together. The current is passed for a suitable length of time controlled by a timer. The various types of resistance welding processes may be divided into the following four main groups : **(*i*)** spot welding **(*ii*)** seam welding **(*iii*)** projection welding and **(*iv*)** butt welding which could be further subdivided into flash welding, upset welding and stud welding etc.

Advantages

Some of the advantages of resistance welding are as under:

1. Heat is localized where required
2. Welding action is rapid
3. No filler material is needed
4. Requires comparatively lesser skill
5. Is suitable for large quantity production
6. Both similar and dissimilar metals can be welded
7. Parent metal is not harmed
8. Difficult shapes and sections can be welded.

Only disadvantages are with regard to high initial as well as maintenance cost.

Spot Welding

It is a form of resistance welding in which the two surfaces are joined by spots of fused metal caused by fused metal between suitable electrodes under pressure.

The process depends on two factors :

1. Resistance heating of small portions of the two workpieces to plastic state and

2. Application of forging pressure for welding the two workpieces.

Heat produced is $H = I^2 Rt\ J$. The resistance R is made up of **(*i*)** resistance of the electrodes and metals themselves **(*ii*)** contact resistance between electrodes and workpieces and **(*iii*)** contact resistance between the two workpieces. Generally, contact resistance between the two workpieces is the greatest.

Fig. 1.49 Spot Welding Machine

As shown in 1.50 (*b*), mechanical pressure is applied by the tips of the two electrodes. In fact, these electrodes not only provide the forging pressure but also carry the welding current and concentrate the welding heat on the weld spot directly below them. Figure 1.50(*a*) shows diagrammatically the basic parts of a modern spot welding. It consists of a step-down transformer which can supply huge currents (upto 5,000 A) for short duration of time.

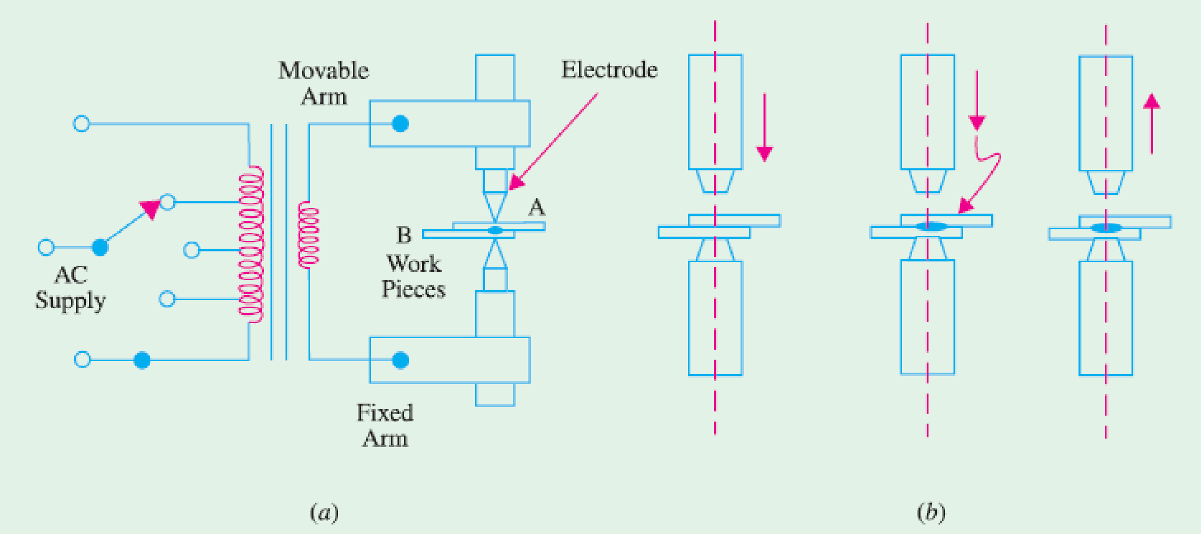


Fig. 1.50

The lower arm is fixed whereas the upper one is movable. The electrodes are made of low-resistance, hard-copper alloy and are either air cooled or butt-cooled by water circulating through the rifled drillings in the electrode. Pointed electrodes [Fig. 1.51 (*a*)] are used for ferrous materials whereas domed electrodes [Fig. 1.51 (*b*)] are used for non-ferrous materials. Flat domes are used when spot-welding deformation is not desired. The weld size is determined by the diameter of the electrode.

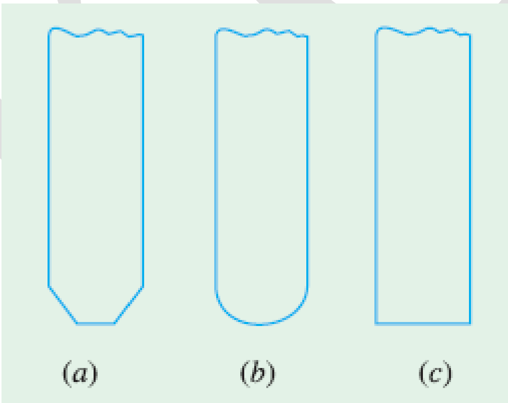


Fig. 1.51

The welding machine is cycled in order to produce the required heat timed to coincide with the pressure exerted by the electrodes as shown in figure 1.50 (*a*). As the movable electrode comes down and presses the two work pieces *A* and *B* together, current is passed through the assembly. The metals under the pressure zone get heated upto about 950°C and fuse together. As they fuse, their resistance is reduced to zero, hence there is a surge of current. This surge is made to switch off the welding current automatically. In motor-driven machines, speeds of 300 strokes minute are common. Spot welders are of two different types. One is a stationary welder which is available in different sizes. The other has a stationary transformer but the electrodes are in a gun form.

Electric resistance spot welding is probably the best known and most widely-used because of its low cost, speed and dependability. It can be easily performed by even a semiskilled operator. This process has a fast welding rate and quick set-up time apart from having low unit cost per weld.

Spot welding is used for galvanized, tinned and lead coated sheets and mild steel sheet work. This technique is also applied to non-ferrous materials such as brass, aluminium, nickel and bronze etc.

Seam Welding

The seam welder differs from ordinary spot welder only in respect of its electrodes which are of disc or roller shape as shown in figure 1.52 (*a*). These copper wheels are power driven and rotate whilst gripping the work. The current is so applied through the wheels that the weld spots either overlap as in figure 1.52 (*b*) or are made at regular intervals as in figure 1.52 (*c*). The continuous or overlapped seam weld is also called ***stitch weld*** whereas the other is called roll weld.

Seam welding is confined to welding of thin materials ranging in thickness from 2 mm to 5 mm. It is also restricted to metals having low hardenability rating such as hot-rolled grades of low alloy steels. Stitch welding is commonly used for long water-tight and gas-tight joints. Roll welding is used for simple joints which are not water-tight or gas-tight. Seam welds are usually tested by pillow test.

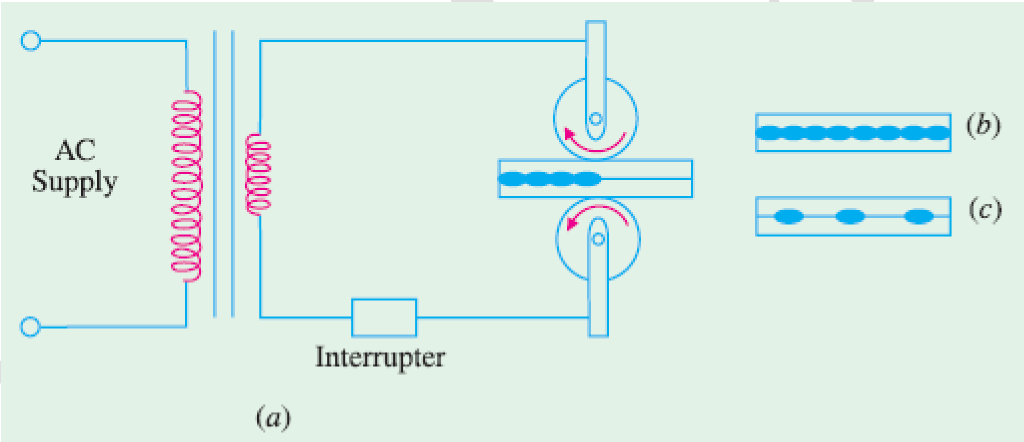


Fig. 1.52

Projection Welding

It can be regarded as a mass-production form of spot welding. Technically, it is a cross between spot welding and butt welding. It uses the same equipment as spot welding. However, in this process, large-diameter flat electrodes (also called platens) are used. This welding process derives its name from the fact that, prior to welding, projections are raised on the surfaces to be welded [Fig. 1.53 (*a*)]. As seen, the upper and lower platens are connected across the secondary of a step-down transformer and are large enough to cover all the projections to be welded at one stroke of the machine. When platen *A* touches the workpiece, welding current flows ***through each projection***.

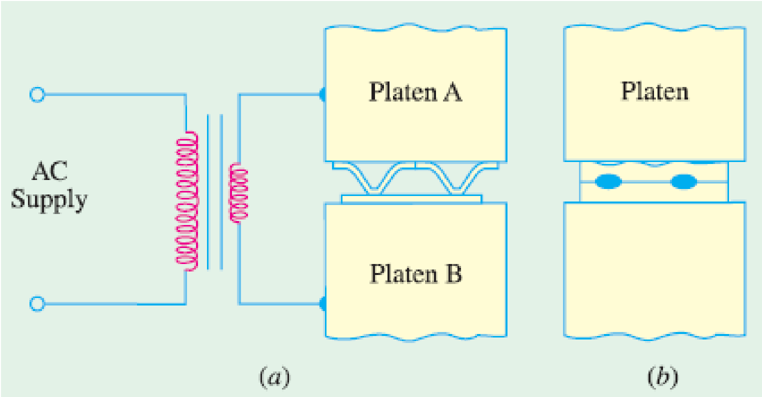


Fig. 1.53

The welding process is started by first lowering the upper platen *A* on to the work-piece and then applying mechanical pressure to ensure correctly-forged welds. Soon after, welding current is switched on as in spot welding. As projection areas heat up, they collapse and union takes place at all projections simultaneously [Fig. 1.53 (*b*)].

It is seen that projections serve many purposes:

1. They increase the welding resistance of the material locally.
2. They accurately locate the positions of the welds.
3. They speed up the welding process by making it possible to perform several small welds simultaneously.
4. They reduce the amount of current and pressure needed to form a good bond between two surfaces.
5. They prolong the life of the electrode considerably because the metal itself controls the heat produced.

Projection welding is used extensively by auto manufactures for joining nuts, bolts and studs to steel plates in car bodies. This process is especially suitable for metals like brass, aluminium and copper etc. mainly due to their high thermal conductivity.

A variation of projection welding is the metal fibre welding which uses a metal fibre rather than a projection point (Fig. 1.54). This metal fibre is generally a felt material. Instead of projections, tiny elements of this felt material are placed between the two metals which are then projection-welded in the usual way.

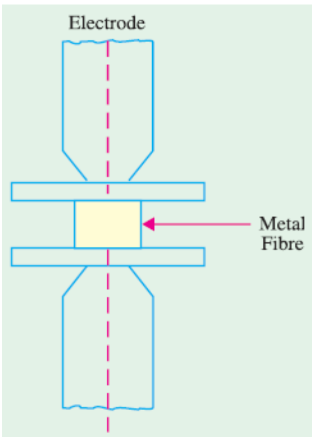


Fig. 1.54

Butt Welding

In this case, the two workpieces are brought into contact end-to-end and the butted ends are heated by passing a heavy current through the joint. As in other forms of resistance welding, the weld heat is produced mainly by the electrical resistance of the joint faces. In this case, however, the electrodes are in the form of powerful vice clamps which hold the work-pieces and also convey the forging pressure to the joint [Fig. 1.55].

This process is useful where parts have to be joined end-to-end or edge-to-edge. *i.e*. for welding pipes, wires and rods. It is also employed for making continuous lengths of chain.

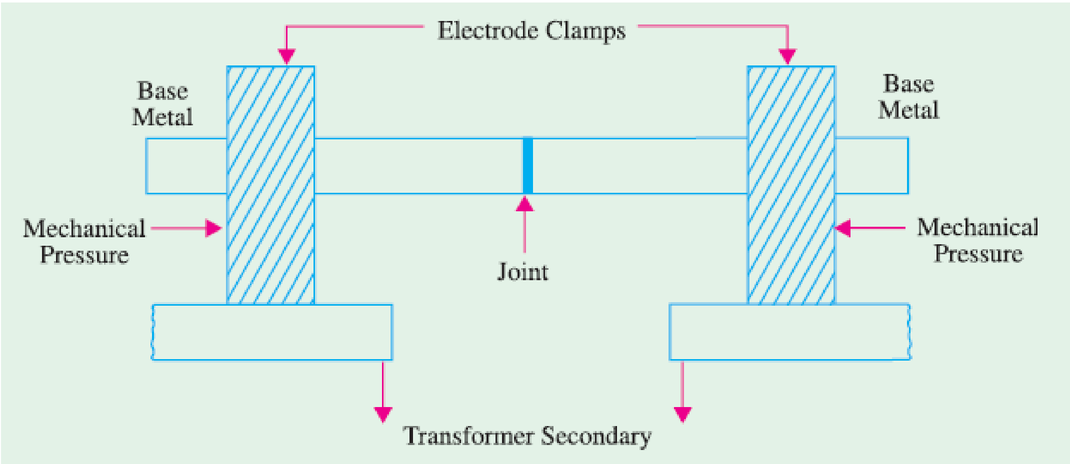


Fig. 1.55

Flash Butt Welding

It is also called by the simple name of ***flash welding***. It is similar to butt welding but with the difference that here current is applied when ends of the two metal pieces are quite close to each other ***but do not touch intimately***. Hence, an arc or flash is set up between them which supplies the necessary welding heat. As seen, in the process heat is applied ***before*** the two parts are pressed together.

As shown in figure 1.56 (*a*), the workpieces to be welded are clamped into specially designed electrodes one of which is fixed whereas the other is movable. After the flash has melted their faces, current is cut off and the movable platen applies the forging pressure to form a fusion weld. As shown in figure 1.56 (*b*), there is increase in the size of the weld zone because of the pressure which forces the soft ends together.

Advantages

1. Even rough or irregular ends can be flash-welded. There is no need to level them by matching and grinding because all irregularities are burnt away during flash period.
2. It is much quicker than butt welding.
3. It uses considerably less current than butt welding.
4. One of its major advantages is that dissimilar metals with different welding temperatures can be flash-welded.

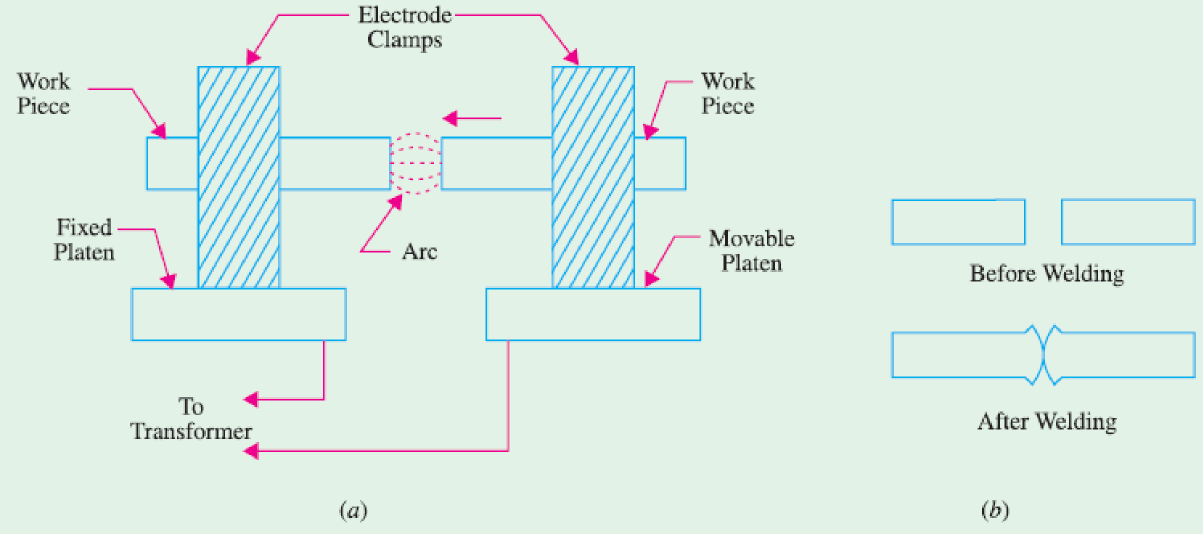


Fig. 1.56

Applications

1. To assemble rods, bars, tubings, sheets and most ferrous metals.
2. In the production of wheel rims for automobiles and bicycles.
3. Foe welding tubular parts such as automobile break cross-shafts.
4. For welding tube coils for refrigeration plants etc.

Upset Welding

In this process, ***no flash is allowed to occur*** between the two pieces of the metals to be welded.

When the two base metals are brought together to a single interface, heavy current is passed between them which heats them up. After their temperature reaches a value of about 950°C, the two pieces of base metal are pressed together more firmly. This pressing together is called ***upsetting***. This upsetting takes place ***while current is flowing and continues even after current is switched off***. This upsetting action mixes the two metals homogeneously while pushing out many atmospheric impurities.

Stud Welding

(a) Basic Principle

It is similar to flash welding because it incorporates a method of drawing an arc between the stud (a rod) and the surface of the base metal. Then, the two molten surfaces are brought together under pressure to form a weld. Stud welding eliminates the need for drilling holes in the main structure.

(b) Welding Equipment

The stud welding equipment consists of a stud welding gun, a d.c. power supply capable of giving currents upto 400 A, a device to control current and studs and ferrules which are used not only as arc shields but also as containing walls for the molten metal.

(c) Applications

It is a low-cost method of fastening extensions (studs) to a metal surface. Most of the ferrous and non-ferrous metals can be stud-welded successfully. Ferrous metals include stainless steel, carbon steel and low-alloy steel. Non-ferrous metals include aluminum, lead-free brass, bronze and chrome plated metals.

Stud welding finds application in the installations of conduit pipe hangers, planking and corrugated roofings.

This process is also used extensively in shipbuilding, railroad and automotive industries.

Plasma Arc Welding

(a) Basic Principle

It consists of a high-current electronic arc which is forced through a small hole in a water-coled metallic nozzle [Fig. 1.57 (*a*)]. The plasma gas itself is used to protect the nozzle from the extreme heat of the arc. The plasma arc is shielded by inert gases like argon and helium which are pumped through an extra passageway within the nozzle of the plasma torch. As seen, plasma arc consists of electronic arc, plasma gas and gases used to shield the jet column. The idea of using the nozzle is to constrict the arc thereby increasing its pressure. Collision of high-energy electrons with gas molecules produces the plasma which is swept through the nozzle and forms the current path between the electrode and the workpiece. Plasma jet torches have temperature capability of about 35,000°C.

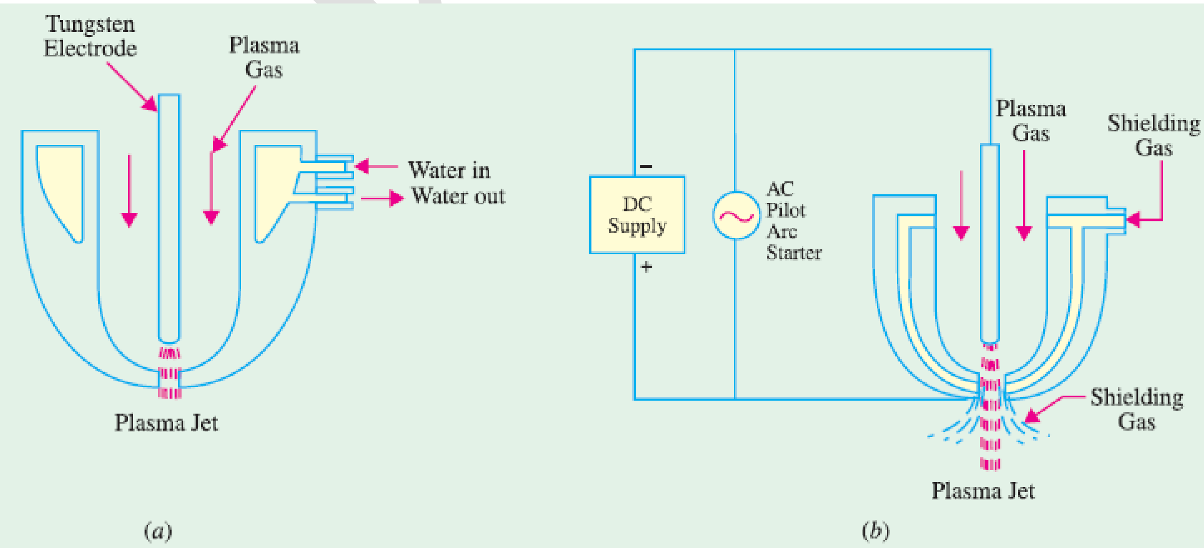


Fig. 1.57

(b) Electrodes

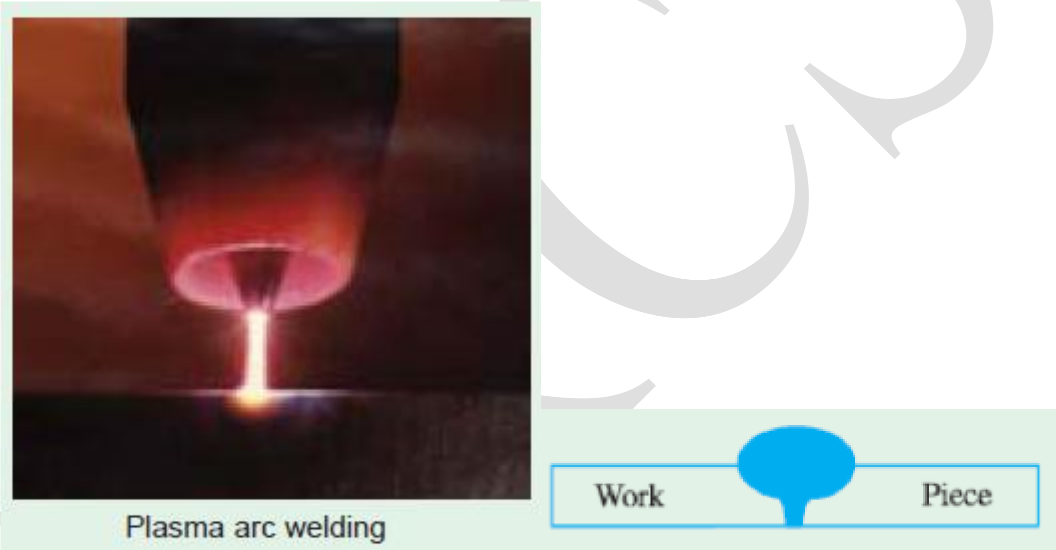
For stainless steel welding and most other metals, straight polarity tungsten electrodes are used. But for aluminium welding, reverse polarity water-cooled copper electrodes are used.

(c) Power Supply

Plasma arc welding requires d.c. power supply which could be provided either by a motor-generator set or transformer-rectifier combination. The latter is preferred because it produces better arc stability. The d.c. supply should have an open-circuit voltage of about 70V and drooping voltage-ampere characteristics. A high-frequency pilot arc circuit is employed to start the arc [Fig. 1.57 (*b*)].

(d) Method of Welding

Welding with plasma arc jet is done by a process called keyhole method. As the plasma jet strikes the surface of the workpiece, it burns a hole through it. As the torch progresses along the work-piece, this hole also progresses along with but is filled up by the molten metal as it moves along. Obviously, 100 percent penetration is achieved in this method of welding. Since plasma jet melts a large surface area of the base metal, it produces a weld bead of wineglass design as shown in figure 1.58. The shape of the bead can be changed by changing the tip of the nozzle of the torch. Practically, all welding is done mechanically.



Plasma arc welding

Fig. 1.58

(e) Applications

1. Plasma arc welding process has many aerospace applications.
2. It is used for welding of reactive metals and thin materials.
3. It is capable of welding high-carbon steel, stainless steel, maraging steel, copper and copper alloys, brass alloys, aluminium and titanium.
4. It is also used for metal spraying.
5. It can be modified for metal cutting purposes. It has been used for cutting aluminium, carbon steel, stainless steel and other hard-to-cut steels. It can produce high-quality drossfree aluminium cuts 15 cm deep.

(*f*) Disadvantages

1. Since it uses more electrical equipment, it has higher electrical hazards.
2. It produces ***ultra-violet and infra-red*** radiations necessitating the use of tinted lenses.

3. It produces high-pitched noise (100 dB) which makes it necessary for the operator to use ear plugs.

Electroslag Welding

(a) General

It is a metal-arc welding process and may be considered as a further development of submerged-arc welding.

This process is used for welding joints of thick sections of ferrous metals in a single pass and without any special joint preparation. Theoretically, there is no upper limit to the thickness of the weld bead. It is usually a vertical uphill process.

It is called ***electroslag*** process because heat is generated by passing current through the molten slag which floats over the top of the metal.

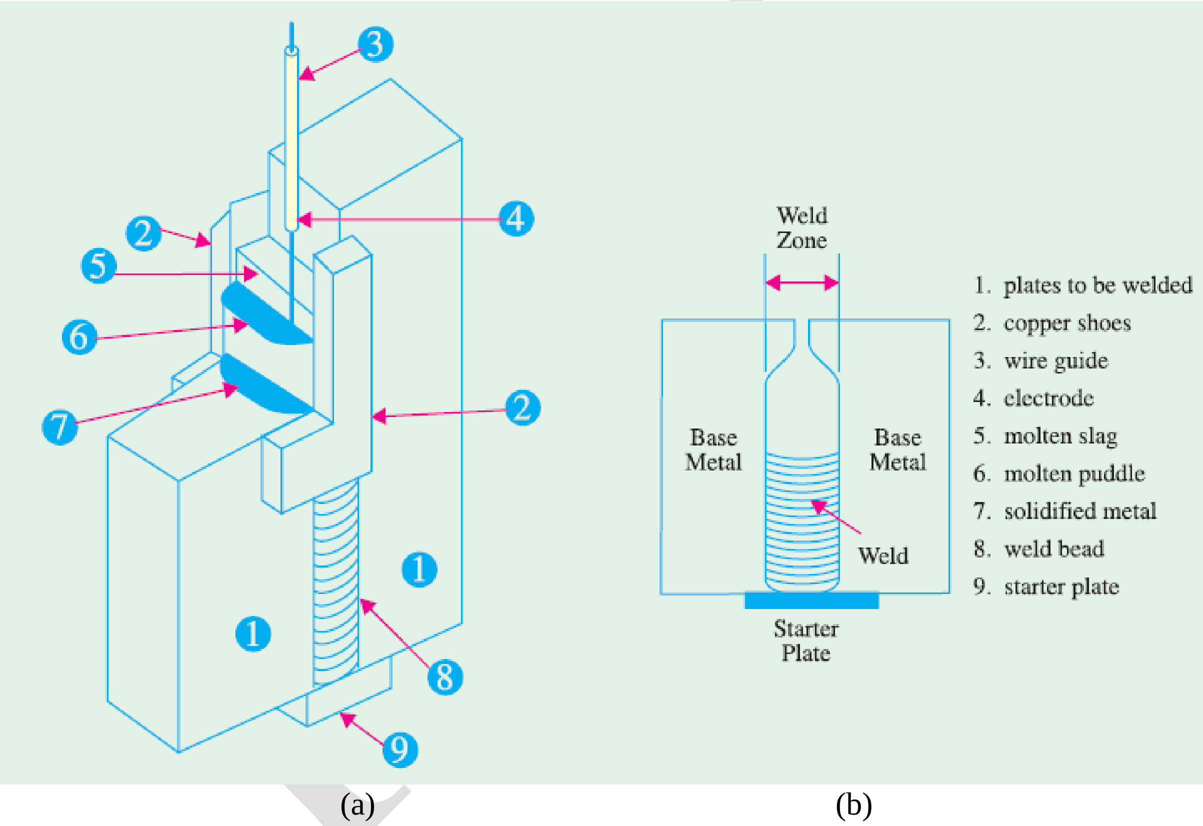


Fig. 1.59

(b) Welding Equipment

As shown in Fig. 1.59, two water-cooled copper shoes (or dams) are placed on either side of the joint to be welded for the purpose of confining the molten metal in the joint area. The electrode is fed into the weld joint almost vertically from special wire guides. There is a mechanical device which raises the shoes and wire-feed mechanism as the weld continues upwards till it is completed.

An a.c. welding machine has 100 percent duty cycle and which can supply currents upto 1000 A if needed.

(c) Welding Process

The electroslag process is initiated just like submerged arc process by starting an electric arc beneath a layer of granular welding flux. When a sufficient thick layer of hot flux or molten slag is formed, arc action stops and from then onwards, current passes from the electrode to workpiece through the molten slag. At this point, the process becomes truly electroslag welding. A starting plate is used in order to build up proper depth of conductive slag before molten pool comes in contact with the work pieces.

The heat generated by the resistance to the flow of current through the molten slag is sufficient to melt the edges of the workpiece and the filler electrode. The molten base metal and filler metal collect at the bottom of the slag pool forming the weld pool. When weld pool solidfies, weld bead is formed which joins the faces of the base metal as shown in figure 1.59(b).

As welding is continued upwards, flux flows to the top in the form of molten slag and cleanses the impurities from the molten metal. A mechanism raises the equipment as the weld is completed in the uphill vertical position.

(d) Advantages

1. It needs no special joint preparation.
2. It does welding in a single pass rather than in costly multiple passes.
3. There is theoretically no maximum thickness of the plate it can weld.
4. There is also no theoretical upper limit to the thickness of the weld bead. Weld beads upto 400 mm thick have been performed with the presently available equipment.
5. This process requires less electrical power per kg of deposited metal than either the submerged arc welding process or the shield arc process.
6. It has high deposit rate of upto 20 kg of weld metal per hour.
7. It has lower flux consumption.
8. Due to uniform heating of the weld area, distortion and residual stresses are reduced to the minimal amounts.

However, for electroslag welding, it is necessary to have only a square butt joint or a square edge on the plates to be welded.

(*e*) Applications

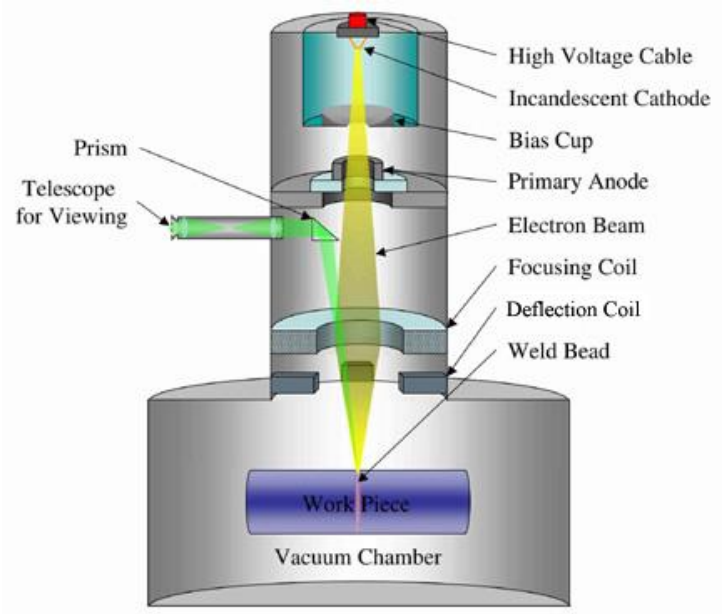
It is commonly used in the fabrication of large vessels and tanks. Low-carbon steels produce excellent welding properties with this process.

Electro Gas Welding

This process works on the same basic principle as the electroslag process but has certain additional features of submerged arc welding. Unlike electroslag process, the electrogas process uses an inert gas for shielding the weld from oxidation and there is a continuous arc (as in submerged arc process) to heat the weld pool.

Electron Beam Welding

In this process, welding operation is performed in a vacuum chamber with the help of a sharply focussed beam of high-velocity electrons. The electrons after being emitted from a suitable electrode are accelerated by the high anode voltage and are then focussed into a fine beam which is finally directed to the workpiece. Obviously, this process needs no electrodes. The electron beam produces intense local heat which can melt not only the metal but can even boil it. A properly-focussed electron beam can completely penetrate through the base metal thereby creating a small hole whose walls are molten. As the beam moves along the joint, it melt the material coming in contact with it. The molten metal flows back to the previously-melted hole where it fuses to make a perfect weld for the entire depth of penetration.




Electron beam welding facility

Fig. 1.60 Electron Beam Welding

Electron-beam welding has following Advantages:

1. It produces deep penetration with little distortion.
2. Its input power is small as compared to other electrical welding devices.
3. Electron-beam weld is much narrower than the fusion weld.
4. It is especially suitable for reactive metals which become contaminated when exposed to air because this process is carried out in vacuum.
5. It completely eliminates the contamination of the weld zone and the weld bead because operation is performed in a vacuum chamber.
6. It is especially suited to the welding of beryllium which is being widely used in the fabrication of industrial and aerospace components.
7. Its high deposition rate produces welds of excellent quality with only a single pass.
8. It is the only process which can join high temperature metals such as columbium.

At present, its only serious limitations are that it is extremely expensive and is not available in portable form. However, recently a non-vacuum electron-beam welder has been developed.

Laser Welding

It uses an extremely concentrated beam of coherent monochromatic light *i.e.* light of only one colour (or wavelength). It concentrates tremendous amount of energy on a very small area of the workpiece to produce fusion. It uses solid laser (ruby, saphire), gas laser (CO_2) and semiconductor laser. Both the gas laser and solid laser need capacitor storage to store energy for later injection into the flash tube which produces the required laser beam.

The gas laser welding equipment consists of ***(i)*** capacitor bank for energy storage ***(ii)*** a triggering device ***(iii)*** a flash tube that is wrapped with wire ***(iv)*** lasing material ***(v)*** focussing lens and ***(vi)*** a worktable that can rotate in the three *X*, *Y* and *Z* directions.

When triggered, the capacitor bank supplies electrical energy to the flash tube through the wire. This energy is then converted into short-duration beam of laser light which is pin-pointed on the workpiece as shown in figure 1.61. Fusion takes place immediately and weld is completed fast.

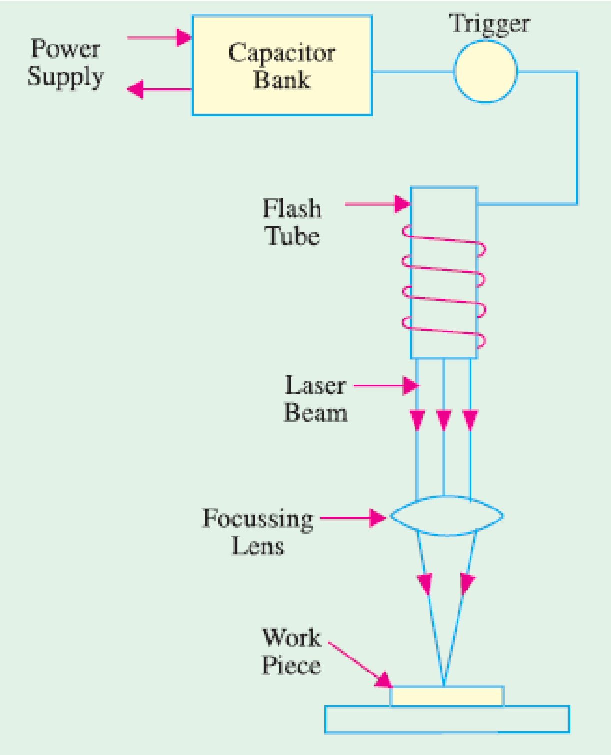


Fig. 1.61

Since duration of laser weld beam is very short (2 ms or so), two basic welding methods have been adopted. In the first method, the workpiece is moved so fast that the entire joint is welded in a single burst of the light. The other method uses a number of pulses one after the other to form the weld joint similar to that formed in electric resistance seam welding.

Laser welding is used in the aircraft and electronic industries for lighter gauge metals. Some of the advantages of laser welding process are as follows:

1. It does not require any electrode.
2. It can make welds with high degree of precision and on materials as thin as 0.025 mm.
3. It does not heat the workpiece except at one point. In fact, heat-affected zone is virtually nonexistent.
4. Liquidus is reached only at the point of fusion.
5. It can produce glass-to-metal seals as in the construction of klystron tubes.
6. Since laser beam is small in size and quick in action, it keeps the weld zone uncontaminated.
7. It can weld dissimilar metals with widely varying physical properties.
8. It produces minimal thermal distortion and shrinkage because area of heat-affected zone is

the minimum possible.

9. It can easily bond refractory materials like molybdenum, titanium and tantalium etc.

However, the major disadvantage of this process is its slow welding speed. Moreover, it is limited to welding with thin metals only.

Questions Bank

1. Explain the properties of good heating element?
2. Define welding, compare resistance and arc welding?
3. A 27kW, 3 phase 400V resistance oven is to employ nickel-chrome strip 0.25mm thick for the three star connected heating elements. If the temperature of the strip is to be 1000^0C and that of the charge to be 600^0C, estimate suitable width for the strip. Assume emissivity = 0.9 and radiating efficiency to be 0.5 and resistivity of the strip material is $101.6\text{x}10^{-8}$ Ω-m.
4. Explain the principles of Induction heating
5. With neat sketch explain the working of a Vertical core type induction furnace.
6. Calculate the efficiency of a high frequency induction furnace which takes 10minutes to melt 1.8kg of aluminium. The input to the furnace is 4.8kW and initial temperature 15^0C. Specific heat of aluminium = 0.88kJ/kg ^{0}C. Melting point of aluminium = 660^0C. Latent heat of fusion of aluminium = 32kJ/kg. Assume 1kJ = $2.78\text{x}10^{-4}$kWh.
7. Explain clearly resistance and arc heating?
8. With neat figure, explain vertical core type induction furnace.
9. A cubic water tank has surface area of 6m2 and is filled to 90% capacity six times daily. The water is heated from 200C to 650C. The loss per square metre of tank surface per 10C temperature difference is 6.3 W. Find the loading kW and the efficiency of the tank. Assume specific heat of water = 4200J/kg/^{0}C and 1kWh = 3.6 MJ.
10. What are the requirements of good welding?
11. Write classification of electric welding.
12. What are the advantages of resistance welding? Explain clearly Butt welding along with its application.
13. Discuss methods of temperature control of resistance oven.
14. Discuss the following applications in dielectric heating. i) Heating of raw plastics ii) Gluing of wood.
15. A cubic water tank has surface area of 5.4m^2 and is filled to 92% capacity five times daily. The water is heated from 15^0C to 60^0C. The losses per square meter of tank per 1^0C temperature difference are 5.9 W. Calculate i) Loading in kW ii) Efficiency of tank. Assume specific heat of water = 4.186kJ/kg/^{0}C
16. Define the term welding. What is resistance welding? What are its limitations?
17. Compare A.C and D.C welding?

UNIT-2

ELECTROLYTIC PROCESS

Definition of Electrolysis

An electrolyte is such a chemical that's atoms are normally closely bonded together but when it is dissolved in water, its molecules split up into positive and negative ions. The positively charged ions are referred as cat ions whereas negatively charged ions are referred as anions. Both cat ions and anions move freely in the solution.

Electrolyte: An electrolyte is a substance that produces an electrically conducting solution when dissolved in a polar solvent, such as water. The dissolved electrolyte separates into cations and anions, which disperse uniformly through the solvent.

Principle of Electrolysis

As discussed in the definition of electrolyte, whenever any electrolyte gets dissolved in water, its molecules split into cat ions and anions moving freely in the electrolytic solution. Now two metal rods are immersed in the solution and an electrical potential difference applied between the rods externally preferably by a battery. These partly immersed rods are technically referred as electrodes. The electrode connected with negative terminal of the battery is known as cathode and the electrode connected with positive terminal of the battery is known as anode. The freely moving positively charged cat ions are attracted by cathode and negatively charged anions are attracted by anode. In cathode, the positive cat ions take electrons from negative cathode and in anode, negative anions give electrons to the positive anode. For continually taking and giving electrons in cathode and anode respectively, there must be flow of electrons in the external circuit of the electrolytic. That means, electric current continues to circulate around the closed loop created by battery, electrolytic and electrodes. This is the most basic **principle of electrolysis**

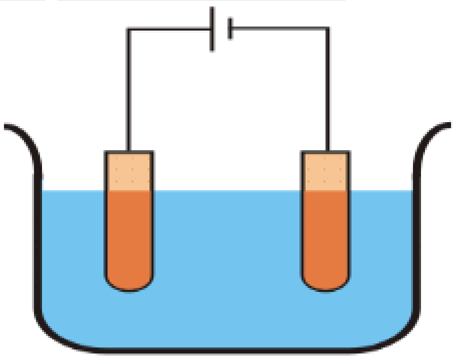


Fig: 2.1 Electrolysis process

Electrolysis of Copper Sulphate

Whenever copper sulphate or $CuSO_4$ is added to water, it gets dissolved in the water. As the $CuSO_4$ is an electrolyte, it splits into Cu^{++} (cat ion) and SO_4^{--} (anion) ions and move freely in the solution. Now if two copper electrodes are immersed in that solution, the Cu^{++} ions (cat ion) will be attracted towards cathode i.e. the electrode connected to the negative terminal of the battery. On reaching on the cathode, each Cu^{++} ion will take electrons from it and becomes neutral copper atoms. Similarly the $SO4^{--}$ (anion) ions will be attracted by anode i.e. the electrode connected to the positive terminal of the battery. So SO_4^{--} ions will move towards anode where they give up two electrons and become SO_4 radical but since SO_4 radical cannot exist in the electrical neutral state, it will attack copper anode and will form copper sulphate. If during **electrolysis of copper sulphate**, we use carbon electrode instead of copper or other metal electrodes, then **electrolysis** reactions will be little bit different. Actually SO_4 cannot react with carbon and in this case the SO_4 will react with water of the solution and will form sulphuric acid and liberate oxygen.

$$2SO_4 + 2H_2O \rightarrow 2H_2SO_4 + O_2$$

The process described above is known as **electrolysis**. In the above process, after taking electrons the neutral copper atoms get deposited on the cathode. At the same time, SO_4 reacts with copper anode and becomes $CuSO_4$ but in water it cannot exist as single molecules instead of that $CuSO_4$ will split into Cu^{++}, SO_4^{--} and dissolve in water. So it can be concluded that, during electrolysis of copper sulphate with copper electrodes, copper is deposited on cathode and same amount of copper is removed from anode.

Faraday's Laws of Electrolysis

Before understanding **Faraday's laws of electrolysis**, we have to recall the process of electrolysis of a metal sulphate.

Whenever an electrolyte like metal sulphate is diluted in water, its molecules split into positive and negative ions. The positive ions or metal ions move to the electrodes connected with negative terminal of the battery where these positive ions take electrons from it, become pure metal atom and get deposited on the electrode. Whereas negative ions or sulphions move to the electrode connected with positive terminal of the battery where these negative ions give up their extra electrons and become SO_4 radical. Since SO_4 cannot exist in electrically neutral state, it will attack metallic positive electrode and form metallic sulphate which will again dissolve in the water. **Faraday's laws of electrolysis** combine two laws and these are,

Faraday's First Law of Electrolysis

From the brief explanation above, it is clear that the flow of current through the external battery circuit fully depends upon how many electrons get transferred from negative electrode or cathode to positive metallic ion or cat ions. If the cat ions have valency of two like Cu++ then for every cat ion, there would be two electrons transferred from cathode to cat ion. We know that every electron has negative electrical charge -1.602×10^{-19} Coulombs and say it is - e. So for disposition of every Cu atom on the cathode, there would be - 2.e charge transfers from cathode to cat ion. Now say for t time there would be total n number of copper atoms deposited on the cathode, so total charge transferred, would be - 2.n.e Coulombs. Mass m of the deposited copper is obviously function of number of atoms deposited. So, it can be concluded that the mass of the deposited copper is directly proportional to the quantity of electrical charge that passes through the electrolyte. Hence mass of deposited copper $m \propto Q$ quantity of electrical charge passes through the electrolyte.

Faraday's First Law of Electrolysis states that only,
According to this law, the chemical deposition due to flow of electric current through an electrolyte is directly proportional to the quantity of electricity (coulombs) passed through it.

i.e. mass of chemical deposition,

$$m \propto Quantity\ of\ electricity,\ Q \Rightarrow m = Z \cdot Q$$

where Z is a constant of proportionality and is known as electrochemical equivalent of the substance.

If we put Q = 1 coulombs in the above equation, we will get Z = m which implies that electrochemical equivalent of any substance is the amount of the substance deposited on passing of 1 coulomb through its solution. This constant of passing of electrochemical equivalent is generally expressed in terms of milligram per coulomb or kilogram per coulomb.

Faraday's Second Law of Electrolysis

So far we have learned that the mass of the chemical, deposited due to electrolysis is proportional to the quantity of electricity that passes through the electrolyte. The mass of the chemical, deposited due to electrolysis is not only proportional to the quantity of electricity passes through the electrolyte, but it also depends upon some other factor. Every substance will have its own atomic weight. So for same number of atoms, different substances will have different masses. Again, how many atoms deposited on the electrodes also depends upon their number of valency. If valency is more, then for same amount of electricity, number of deposited atoms will be less whereas if valency is less, then for same quantity of electricity, more number of atoms to be deposited. So, for same quantity of electricity or charge passes through different electrolytes, the mass of deposited chemical is directly proportional to its atomic weight and inversely proportional to its valency.

Faraday's second law of electrolysis states that, when the same quantity of electricity is passed through several electrolytes, the mass of the substances deposited are proportional to their respective chemical equivalent or equivalent weight.

Applications of Electrolysis:

Electrolytic Refining of Metals

The process of **electrolytic refining of metals** is used to extract impurities from crude metals. Here in this process, a block of crude metal is used as anode, a diluted salt of that metal is used as electrolyte and plates of that pure metal is used as cathode.

Electrolytic Refining of Copper

For understanding the process of **electrolytic refining of metals**, we will discuss about an example of **electrolytic refining of copper**. Copper extracted from its ore, known as blister copper, is 98 to 99 % pure but it can easily be made up to 99.95% pure for electrical application by the process of **electrorefining**.

In this process of electrolysis, we use a block of impure copper as anode or positive electrode, copper sulphate acidified with sulphuric acid, electrolyte and pure copper plates coated with graphite, as cathode or negative electrode.

The copper sulphate splits into positive copper ion (Cu^{++}) and negative sulphate ion (SO_4^{--}). The positive copper ion (Cu^{++}) or cations will move towards negative electrode made of pure copper where it takes electrons from cathode, becomes Cu atom and is deposited on the graphite surface of the cathode.

On the other hand, the SO_4^{--} will move towards positive electrode or anode where it will receive electrons from anode and become radical SO_4 but as radical SO_4 cannot exist alone, it will attack copper of anode and form CuSO4. This $CuSO_4$ will then dissolve and split in the solution as positive copper ion (Cu^{++}) and negative sulphate ion (SO_4^{--}). These positive copper ions (Cu^{++}) will then move towards negative electrode where it takes electrons from cathode, become Cu atoms and are deposited on the graphite surface of the cathode. In this way, the copper of impure crude will be transferred and deposited on the graphite surface of the cathode. The metallic impurities of anode are also merged with SO4, form metallic sulphate and dissolve in the electrolyte solution. The impurities like silver and gold, which are not effected by sulfuric acid-copper sulphate solution, will settle down as the anode sludge or mud. At a regular interval of electrolytic refining of copper, the deposited copper is stripped out from the cathode and anode is replaced by a new block of crude copper.

Electroplating

The process of **electroplating** is theoretically same as electro refining - only difference is that, in place of graphite coated cathode we have to place an object on which the **electroplating** has to be done. Let's take an example of brass key which is to be copper-platted by using **copper electroplating**.

Copper Electroplating

We have already stated that copper sulphate splits into positive copper ion (Cu^{++}) and negative sulphate ion (SO_4^{--}) in its solution. For **copper electroplating**, we use copper sulphate solution as electrolyte, pure copper as anode and an object (a brass key) as cathode. The pure copper rod is connected with positive terminal and the brass key is connected with negative terminal of a battery. While these copper rod and key are immersed into copper-sulphate solution, the copper rod will behave as anode and the key will behave as cathode. As the cathode or the brass key is connected with negative terminal of battery, it will attract the positive cations or Cu^{++} ions and on reaching of Cu^{++} ions on the surface of the brass key, they will receive electrons from it, become neutral copper atom and are about to be deposited on the surface of the brass key as uniform layer. The sulphate or SO_4^{--} ions move to the anode and extract copper from it into the solution as mentioned in the process of electro-refining. For proper and uniform copper plating, the object (here it is brass key) is being rotated slowly into the solution.

Electroforming

Reproduction of objects by electro-deposition on some sort of mould is known as **electroforming**. This is another very useful example among many applications of electrolysis. For that, first we have to take the impression of objects on wax or on other wax like material. The surface of the wax mold which bears exact impression of the object, is coated with graphite powder in order to make it conducting. Then the mold is dipped into the electrolyte solution as cathode. During electrolysis process, the electrolyte metal will be deposited on the graphite coated impressed surface of the mold. After obtaining a layer of desired thickness, the article is removed and the wax is melted to get the reproduced object in form of metal shell.

A popular use of **electroforming** is reproduction of gramophone record dices. The original recording is done on a record of wax composition. This wax mold is then coated with gold powder to make it conducting. Then this mold is dipped into a blue vitriol electrolyte as cathode. The solution is kept saturated by using a copper anode. The copper electroforming on the wax mold produces master plate which is used to stamp a large number of shellac discs.

Factors affecting Electro deposition Process

Electrophoretic deposition (**EPD**):

It is a term for a broad range of industrial processes which includes **electrocoating**, **e-coating**, **cathodic electrodeposition**, **anodic electrodeposition**, and **electrophoretic coating**, or **electrophoretic painting**. A characteristic feature of this process is that colloidal particles suspended in a liquid medium migrate under the influence of an electric field (electrophoresis) and are deposited onto an electrode. All colloidal particles that can be used to form stable suspensions and that can carry a charge can be used in electrophoretic deposition. This includes materials such as polymers, pigments, dyes, ceramics and metals. The process is useful for applying materials to any electrically conductive surface. The materials which are being deposited are the major determining factor in the actual processing conditions and equipment which may be used.

Due to the wide utilization of electrophoretic painting processes in many industries, aqueous EPD is the most common commercially used EPD process. However, non-aqueous electrophoretic deposition applications are known. Applications of non-aqueous EPD are currently being explored for use in the fabrication of electronic components and the production of ceramic coatings. Non-aqueous processes have the advantage of avoiding the electrolysis of water and the oxygen evolution which accompanies electrolysis.

Uses of EPD

This process is industrially used for applying coatings to metal fabricated products. It has been widely used to coat automobile bodies and parts, tractors and heavy equipment, electrical switch gear, appliances, metal furniture, beverage containers, fasteners, and many other industrial products.

EPD processes are often applied for the fabrication of supported titanium dioxide (TiO_2) photocatalysts for water purification applications, using precursor powders which can be immobilised using EPD methods onto various support materials. Thick films produced this way allow cheaper and more rapid synthesis relative to sol-gel thin-films, along with higher levels of photocatalyst surface area.

EPD processed have a number of advantages which have made such methods widely used

1. The process applies coatings which generally have a very uniform coating thickness without porosity.
2. Complex fabricated objects can easily be coated, both inside cavities as well as on the outside surfaces.
3. Relatively high speed of coating.
4. Relatively high purity.
5. Applicability to wide range of materials (metals, ceramics, polymers, etc.)

6. Easy control of the coating composition.
7. The process is normally automated and requires less human labour than other coating processes.
8. Highly efficient utilization of the coating materials result in lower costs relative to other processes.
9. The aqueous process which is commonly used has less risk of fire relative to the solvent-borne coatings that they have replaced.
10. Modern electrophoretic paint products are significantly more environmentally friendly than many other painting technologies.

Process of electrophoretic painting

The overall industrial process of electrophoretic deposition consists of several sub-processes:

1. The object to be coated needs to be prepared for coating. This normally consists of some kind of cleaning process and may include the application of a conversion coating, typically an inorganic phosphate coating.
2. The coating process itself. This normally involves submerging the part into a container or vessel which holds the coating bath or solution and applying direct current electricity through the EPD bath using electrodes. Typically voltages of 25 - 400 volts DC are used in electrocoating or electrophoretic painting applications. The object to be coated is one of the electrodes, and a set of "counter-electrodes" are used to complete the circuit.
3. After deposition, the object is normally rinsed to remove the undeposited bath. The rinsing process may utilize an ultrafilter to dewater a portion of the bath from the coating vessel to be used as rinse material. If an ultrafilter is used, all of the rinsed off materials can be returned to the coating vessel, allowing for high utilization efficiency of the coating materials, as well as reducing the amount of waste discharged into the environment.
4. A baking or curing process is normally used following the rinse. This will crosslink the polymer and allows the coating, which will be porous due to the evolution of gas during the deposition process, to flow out and become smooth and continuous.

Power Supply for electrolytic process

Electrolysis is a method of using a direct electric current (DC) to drive an otherwise non-spontaneous chemical reaction. Electrolysis is commercially highly important as a stage in the separation of elements from naturally occurring sources such as ores using an electrolytic cell. The voltage that is needed for electrolysis to occur is called decomposition potential.

Electrolysis is the passage of a direct electric current through an ionic substance that is either molten or dissolved in a suitable solvent, resulting in chemical reactions at the electrodes and separation of materials.

The main components required to achieve electrolysis are:

An electrolyte: a substance containing free ions which are the carriers of electric current in the electrolyte. If the ions are not mobile, as in a solid salt then electrolysis cannot occur.

A direct current (DC) supply: provides the energy necessary to create or discharge the ions in the electrolyte. Electric current is carried by electrons in the external circuit.

Two electrodes: an electrical conductor which provides the physical interface between the electrical circuit providing the energy and the electrolyte

Electrodes of metal, graphite and semiconductor material are widely used. Choice of suitable electrode depends on chemical reactivity between the electrode and electrolyte and the cost of manufacture.

Energy changes during electrolysis

The amount of electrical energy that must be added equals the change in Gibbs free energy of the reaction plus the losses in the system. The losses can (in theory) be arbitrarily close to zero, so the maximum thermodynamic efficiency equals the enthalpy change divided by the free energy change of the reaction. In most cases, the electric input is larger than the enthalpy change of the reaction, so some energy is released in the form of heat. In some cases, for instance, in the electrolysis of steam into hydrogen and oxygen at high temperature, the opposite is true. Heat is absorbed from the surroundings, and the heating value of the produced hydrogen is higher than the electric input.

Factors affecting Electro deposition process

The factors affecting the electro deposition process or Electrophoretic Deposition (EPD) are as follows:

1. Current density
2. Electrolytic concentration
3. Temperature
4. Nature of electrolyte
5. Addition agents
6. Nature of the metal on which the deposit is to be made
7. Throwing power of electrolyte

These factors are discussed below in details:

1. ***Current density***: At low values of current density the ions are released at a slow rate and the rate of growth of nuclei is more than the rate at which the new nuclei form themselves. At higher values of current density the quality of deposit becomes more uniform.
2. ***Electrolytic concentration***: This is more or less complimentary to the first factor. i.e. by increasing the concentration of the electrolyte, higher current density can be obtained.
3. ***Temperature:*** The temperature of the electrolyte has two contradictory effects. One, at comparatively high temperature there is more diffusion and even at relatively high current density smooth deposits may be produced. Two, the rate of crystal growth increases the possibility of coarse deposits. At moderate temperatures the deposits are good. In chromium plating the temperature is maintained at 35^0C, and in nickel between 50^0C to 60^0C.
4. ***Nature of electrolyte:*** The formation of smooth deposit largely depends upon the nature of electrolyte used. Smooth deposits are obtained from solutions having complex ions.
5. ***Addition agents***: The quality of a deposit is improved by the presence of an addition agent which may be colloidal matter or an organic compound, otherwise the metal deposits in the form of large crystals and the surface becomes rough.
6. ***Nature of the metal on which deposit is to be made***: This factor influences the growth of crystals since it is believed that the operation of crystals is in continuation of these in the base

metal.

7. ***Throwing power***: The throwing power of an electrolyte may be regarded as the quality which produces a uniform deposit on a cathode having an irregular shape.

Questions Bank

1. What do you mean by Electrolysis process?
2. Explain the extraction of metals.
3. Define electro chemical equivalent and energy efficiency.
4. State and explain faradays laws of electrolysis.
5. Explain throwing power and polarisation.
6. What do you mean by electro deposition?
7. Explain the factors affecting electro deposition.

ELECTRIC TRACTION

Introduction

By electric traction is meant locomotion in which the driving (or tractive) force is obtained from electric motors. It is used in electric trains, tramcars, trolley buses and diesel-electric vehicles etc. Electric traction has many advantages as compared to other non-electrical systems of traction including steam traction.

Traction Systems

Broadly speaking, all traction systems may be classified into two categories:

(a) Non-electric traction systems

They do not involve the use of electrical energy at ***any stage***. Examples are: steam engine drive used in railways and internal-combustion-engine drive used for road transport.

(b) Electric traction systems

They involve the use of electric energy at some stage or the other. They may be further subdivided into two groups:

1. First group consists of self-contained vehicles or locomotives. Examples are: battery-electric drive and diesel-electric drive etc.
2. Second group consists of vehicles which receive electric power from a distribution network fed at suitable points from either central power stations or suitably-spaced sub-stations. Examples are: railway electric locomotive fed from overhead ac supply and tramways and trolley buses supplied with dc supply.

Direct Steam Engine Drive

Though losing ground gradually due to various reasons, steam locomotive is still the most widely adopted means of propulsion for railway work. Invariably, the reciprocating engine is employed because

1. It is inherently simple.
2. Connection between its cylinders and the driving wheels is simple.
3. Its speed can be controlled very easily.

However, the steam locomotive suffers from the following disadvantages:

1. Since it is difficult to install a condenser on a locomotive, the steam engine runs non-condensing and, therefore, has a very low thermal efficiency of about 6-8 percent.
2. It has strictly limited overload capacity.
3. It is available for hauling work for about 60% of its working days, the remaining 40% being spent in preparing for service, in maintenance and overhaul.

Diesel-electric Drive

It is a self-contained motive power unit which employs a diesel engine for direct drive of a dc generator. This generator supplies current to traction motors which are geared to the driving axles. In India, diesel locomotives were introduced in 1945 for shunting service on broad gauge (BG) sections and in 1956 for high-speed main-line operations on metre-guage

(MG) sections. It was only in 1958 that Indian Railways went in for extensive main-line dieselisation.

Diesel-electric traction has the following advantages:

1. No modification of existing tracks is required while converting from steam to diesel-electric traction.
2. It provides greater tractive effort as compared to steam engine which results in higher starting acceleration.
3. It is available for hauling for about 90% of its working days.
4. Diesel-electric locomotive is more efficient than a steam locomotive (though less efficient than an electric locomotive).

Disadvantages**:**

1. For same power, diesel-electric locomotive is costlier than either the steam or electric locomotive.
2. Overload capacity is limited because diesel engine is a constant-kW output prime mover.
3. Life of a diesel engine is comparatively shorter.
4. Diesel-electric locomotive is heavier than plain electric locomotive because it carries the main engine, generator and traction motors etc.
5. Regenerative braking cannot be employed though rheostatic braking can be.

Battery-electric Drive

In this case, the vehicle carries secondary batteries which supply current to dc motors used for driving the vehicle. Such a drive is well-suited for shunting in railway yards, for traction in mines, for local delivery of goods in large towns and large industrial plants. They have low maintenance cost and are free from smoke. However, the scope of such vehicles is limited because of the small capacity of the batteries and the necessity of charging them frequently.

The above picture shows a battery run car. Battery run vehicles are seen as alternatives for future transport due to their pollution-free locomotion

Advantages of Electric Traction

As compared to steam traction, electric traction has the following advantages:

1. **Cleanliness.** Since it does not produce any smoke or corrosive fumes, electric traction is most suited for underground and tube railways. Also, it causes no damage to the buildings and other apparatus due to the absence of smoke and flue gases.

2. **Maintenance Cost.** The maintenance cost of an electric locomotive is nearly 50% of that for a steam locomotive. Moreover, the maintenance time is also much less.
3. **Starting Time.** An electric locomotive can be started at a moment's notice whereas a steam locomotive requires about two hours to heat up.
4. **High Starting Torque.** The motors used in electric traction have a very high starting torque. Hence, it is possible to achieve higher accelerations of 1.5 to 2.5 km/h/s as against 0.6 to 0.8 km/h/s in steam traction. As a result, we are able to get the following additional advantages:

(i) high schedule speed
(ii) increased traffic handling capacity
(iii) because of (*i*) and (*ii*) above, less terminal space is required – a factor of great importance in urban areas.

5. **Braking.** It is possible to use regenerative braking in electric traction system. It leads to the following advantages :

(i) about 80% of the energy taken from the supply during ascent is returned to it during descent.
(ii) goods traffic on gradients becomes safer and speedier.
(iii) since mechanical brakes are used to a very small extent, maintenance of brake shoes, wheels, tyres and track rails is considerably reduced because of less wear and tear.

6. **Saving in High Grade Coal.** Steam locomotives use costly high-grade coal which is not so abundant. But electric locomotives can be fed either from hydroelectric stations or pit-head thermal power stations which use cheap low-grade coal. In this way, high-grade coal can be saved for metallurgical purposes.
7. **Lower Centre of Gravity.** Since height of an electric locomotive is much less than that of a steam locomotive, its centre of gravity is comparatively low. This fact enables an electric locomotive to negotiate curves at higher speeds quite safely.
8. **Absence of Unbalanced Forces.** Electric traction has higher coefficient of adhesion since there are no unbalanced forces produced by reciprocating masses as is the case in steam traction. It not only reduces the weight/kW ratio of an electric locomotive but also improves its riding quality in addition to reducing the wear and tear of the track rails.

Disadvantages of Electric Traction

1. The most vital factor against electric traction is the initial high cost of laying out overhead electric supply system. Unless the traffic to be handled is heavy, electric traction becomes uneconomical.
2. Power failure for few minutes can cause traffic dislocation for hours.
3. Communication lines which usually run parallel to the power supply lines suffer from electrical interference. Hence, these communication lines have either to be removed away from the rail track or else underground cables have to be used for the purpose which makes the entire system still more expensive.
4. Electric traction can be used only on those routes which have been electrified. Obviously, this restriction does not apply to steam traction.
5. Provision of a negative booster is essential in the case of electric traction. By avoiding the flow of return currents through earth, it curtails corrosion of underground pipe work and interference with telegraph and telephone circuits.

Systems of Railway Electrification

Presently, following four types of track electrification systems are available:

1. Direct current system—600 V, 750 V, 1500 V, 3000 V

2. Single-phase ac system—15-25 kV, $16\frac{2}{3}$, 25 and 50 Hz

3. Three-phase ac system—3000-3500 V at $16\frac{2}{3}$ Hz

4. Composite system—involving conversion of single-phase ac into 3-phase ac or dc.

Direct Current System

Direct current at 600-750 V is universally employed for tramways in urban areas and for many suburban railways while 1500-3000 V dc is used for main line railways. The current collection is from third rail (or conductor rail) up to 750 V, where large currents are involved and from overhead wire for 1500 V and 3000 V, where small currents are involved. Since in majority of cases, track (or running) rails are used as the return conductor, only one conductor rail is required. Both of these contact systems are fed from substations which are spaced 3 to 5 km for heavy suburban traffic and40-50 km for main lines operating at higher voltages of 1500 V to 3000 V. These sub-stations themselves receive power from 110/132 kV, 3-phase network (or grid). At these substations, this high-voltage 3-phase supply is converted into low-voltage 1-phase supply with the help of Scott connected or V-connected 3-phase transformers. Next, this low ac voltage is converted into the required dc voltage by using suitable rectifiers or converters (like rotary converter, mercury arc, metal or semiconductor rectifiers). These substations are usually automatic and are remote controlled.

The dc supply so obtained is fed via suitable contact system to the traction motors which are either dc series motors for electric locomotive or compound motors for tramway and trolley buses where regenerative braking is desired.

It may be noted that for ***heavy suburban service***, low voltage dc system is undoubtedly superior to 1phase ac system due to the following reasons:

1. DC motors are better suited for frequent and rapid acceleration of heavy trains than ac motors.
2. DC train equipment is lighter, less costly and more efficient than similar ac equipment.
3. When operating under similar service conditions, dc train consumes less energy than a 1-phase ac train.
4. The conductor rail for dc distribution system is less costly, both initially and in maintenance than the high-voltage overhead ac distribution system.
5. DC system causes no electrical interference with overhead communication lines.

The only disadvantage of dc system is the necessity of locating ac/dc conversion sub-stations at relatively short distances apart.

Single-Phase Low-frequency AC System

In this system, ac voltages from 11 to 15 kV at $16\frac{2}{3}$ or 25 Hz are used. If supply is from a generating station exclusively meant for the traction system, there is no difficulty in getting the electric supply of $16\frac{2}{3}$ or 25 Hz. If, however, electric supply is taken from the high voltage transmission lines at 50 Hz, then in addition to step-down transformer, the substation is provided with a frequency converter. The frequency converter equipment consists of a 3-phase synchronous motor which drives a 1-phase alternator having or 25 Hz frequency.

The 15 kV $16\frac{2}{3}$ or 25 Hz supply is fed to the electric locomotors via a single over-head wire (running rail providing the return path).

A step-down transformer carried by the locomotive reduces the 15-kV voltage to 300 - 400 V for feeding the ac series motors. Speed regulation of ac series motors is achieved by applying variable voltage from the tapped secondary of the above transformer. Low-frequency ac supply is used because apart from improving the commutation properties of ac motors, it increases their efficiency and power factor. Moreover, at low frequency, line reactance is less so that line impedance drop and hence line voltage drop is reduced. Because of this reduced line drop, it is feasible to space the substations 50 to 80 k km apart. Another advantage of employing low frequency is that it reduces telephonic interference.

Three-phase Low-frequency AC System

It uses 3-phase induction motors which work on a 3.3 kV$16\frac{2}{3}$ Hz supply. Sub-stations receive power at a very high voltage from 3-phase transmission lines at the usual industrial frequency of 50 Hz. This high voltage is stepped down to 3.3 kV by transformers whereas frequency is reduced from 50 Hz to 23 16 Hz by frequency converters installed at the substations. Obviously, this system employs *two* overhead contact wires, the track rail forming the third phase (of course, this leads to insulation difficulties at the junctions). Induction motors used in the system are quite simple and robust and give trouble-free operation.

They possess the merits of high efficiency and of operating as a generator when driven at speeds above the synchronous speed. Hence, they have the property of automatic regenerative braking during the descent on gradients. However, it may be noted that despite all its advantages, this system has not found much favour and has; in fact, become obsolete because of its certain inherent limitations given below:

1. The overhead contact wire system becomes complicated at crossings and junctions.
2. Constant-speed characteristics of induction motors are not suitable for traction work.
3. Induction motors have speed/torque characteristics similar to dc shunt motors. Hence, they are not suitable for parallel operation because, even with little difference in rotational speeds caused by unequal diameters of the wheels, motors will becomes loaded very unevenly.

Composite System

Such a system incorporates good points of two systems while ignoring their bad points. Two such composite systems presently in use are:

1. 1-phase to 3-phase system also called Kando system
2. 1-phase to dc system.

Kando System

In this system, single-phase 16-kV, 50 Hz supply from the sub-station is picked up by the locomotive through the single overhead contact wire. It is then converted into 3-phase ac supply at the same frequency by means of phase converter equipment carried on the locomotives. This 3-phase supply is then fed to the 3-phase induction motors.

As seen, the complicated overhead two contact wire arrangement of ordinary 3-phase system is replaced by a single wire system. By using silicon controlled rectifier as inverter, it is possible to get variable-frequency 3-phase supply at 1/2 to 9 Hz frequency. At this low frequency, 3-phase motors develop high starting torque without taking excessive current. In view of the above, Kando system is likely to be developed further.

Single-phase AC to DC System

This system combines the advantages of high-voltage ac distribution at industrial frequency with the dc series motors traction. It employs overhead 25-kV, 50-Hz supply which is stepped down by the transformer installed in the locomotive itself. The low-voltage ac supply is then converted into dc supply by the rectifier which is also carried on the locomotive. This dc supply is finally fed to dc series traction motor fitted between the wheels. The system of traction employing 25-kV, 50-Hz, 1-phase ac supply has been adopted for all future track electrification in India.

Advantages of 25-kV, 50-Hz AC System

Advantages of this system of track electrification over other systems particularly the dc system are as under:

1. **Light Overhead Catenary**
Since voltage is high (25 kV), line current for a given traction demand is less. Hence, cross section of the overhead conductors is reduced. Since these small-sized conductors are light, supporting structures and foundations are also light and simple. Of course, high voltage needs higher insulation which increases the cost of overhead equipment (OHE) but the reduction in the size of conductors has an overriding effect.

2. **Less Number of Substations**
Since in the 25-kV system, line current is less, line voltage drop which is mainly due to the resistance of the line is correspondingly less. It improves the voltage regulation of the line which fact makes larger spacing of 50-80 km between sub-stations possible as against 5-15 km with 1500 V dc system and 15-30 km with 3000 V dc sysem. Since the required number of substations along the track is considerably reduced, it leads to substantial saving in the capital expenditure on track electrification.

3. **Flexibility in the Location of Substations**
Larger spacing of substations leads to greater flexibility in the selection of site for their proper location. These substations can be located near the national high-voltage grid which, in our country, fortunately runs close to the main railway routes. The substations are fed from this grid thereby saving the railway administration lot of expenditure for erecting special transmission lines for their substations. On the other hand, in view of closer spacing of dc

substations and their far away location, railway administration has to erect its own transmission lines for taking feed from the national grid to the substations which consequently increases the initial cost of electrification.

4. **Simplicity of Substation Design**
   In ac systems, the substations are simple in design and layout because they do not have to install and maintain rotary converters or rectifiers as in dc systems. They only consist of static transformers along with their associated switchgear and take their power directly from the high-voltage national grid running over the length and breadth of our country. Since such sub-stations are remotely controlled, they have few attending personnel or even may be unattended.

5. **Lower Cost of Fixed Installations**
   The cost of fixed installations is much less for 25 kV ac system as compared to dc system. In fact, cost is in ascending order for 25 kV ac, 3000 V dc and 1500 V dc systems. Consequently, traffic densities for which these systems are economical are also in the ascending order

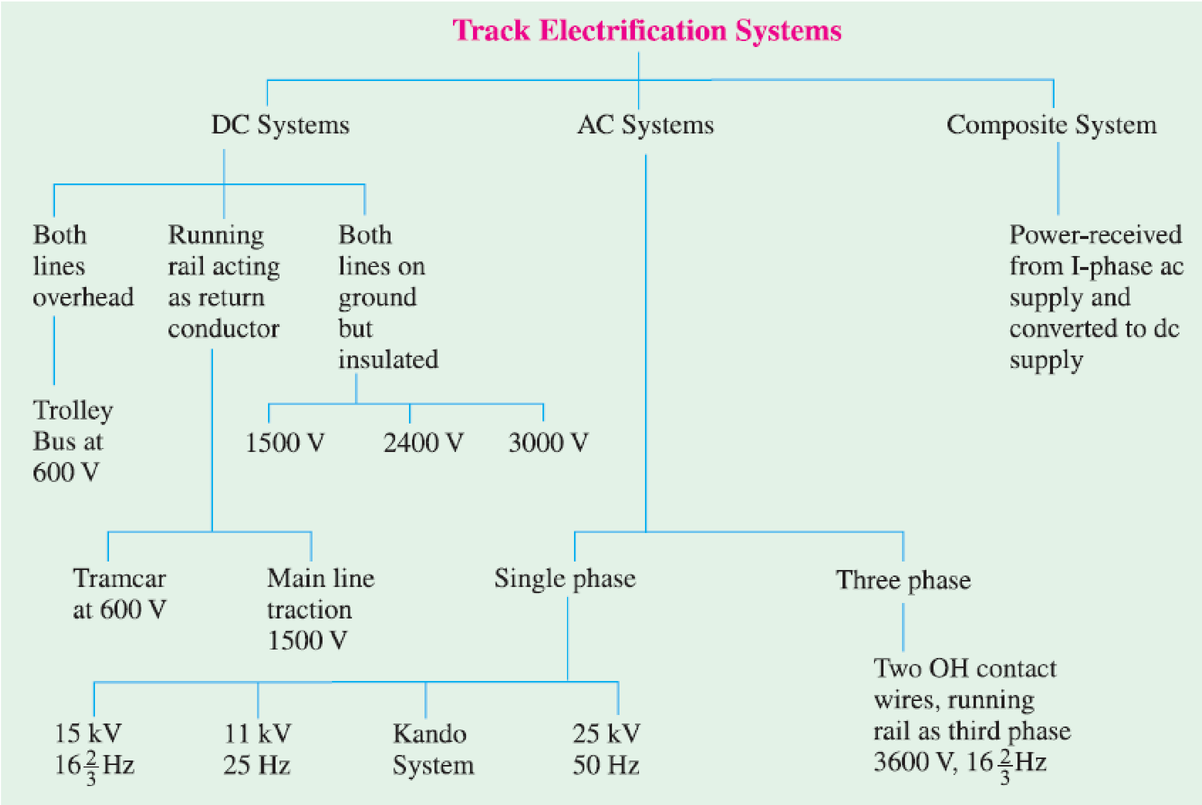
6. **Higher Coefficient of Adhesion**
   The straight dc locomotive has a coefficient of adhesion of about 27% whereas its value for ac rectifier locomotive is nearly 45%. For this reason, a lighter ac locomotive can haul the same load as a heavier straight dc locomotive. Consequently, ac locomotives are capable of achieving higher speeds in coping with heavier traffic.

7. **Higher Starting Efficiency**
   An AC locomotive has higher starting efficiency than a straight dc locomotive. In dc locomotive supply voltage at starting is reduced by means of ohmic resistors but by on-load primary or secondary tap-changer in ac locomotives.

Disadvantages of 25-kV AC System

1. Single-phase ac system produces both current and voltage unbalancing effect on the supply.
2. It produces interference in telecommunication circuits. Fortunately, it is possible at least to minimize both these undesirable effects.

Different track electrification systems are summarised below:



Block Diagram of an AC Locomotive:

The various components of an AC locomotive running on single-phase 25-kV, 50-Hz ac supply are numbered in figure 2.1.

1. OH contact wire
2. Pantograph
3. Circuit breakers
4. On-load tap-changers
5. Transformer
6. Rectifier
7. Smoothing choke
8. DC traction motors.

As seen, power at 25 kV is taken via a pantograph from the overhead contact wire and fed to the step-down transformer in the locomotive. The low ac voltage so obtained is converted into pulsating dc voltage by means of the rectifier. The pulsations in the dc voltage are then removed by the smoothing choke before it is fed to dc series traction motors which are mounted between the wheels. The function of circuit breakers is to immediately disconnect the locomotive from the overhead supply in case of any fault in its electrical system. The on-load tap-changer is used to change the voltage across the motors and hence regulate their speed.

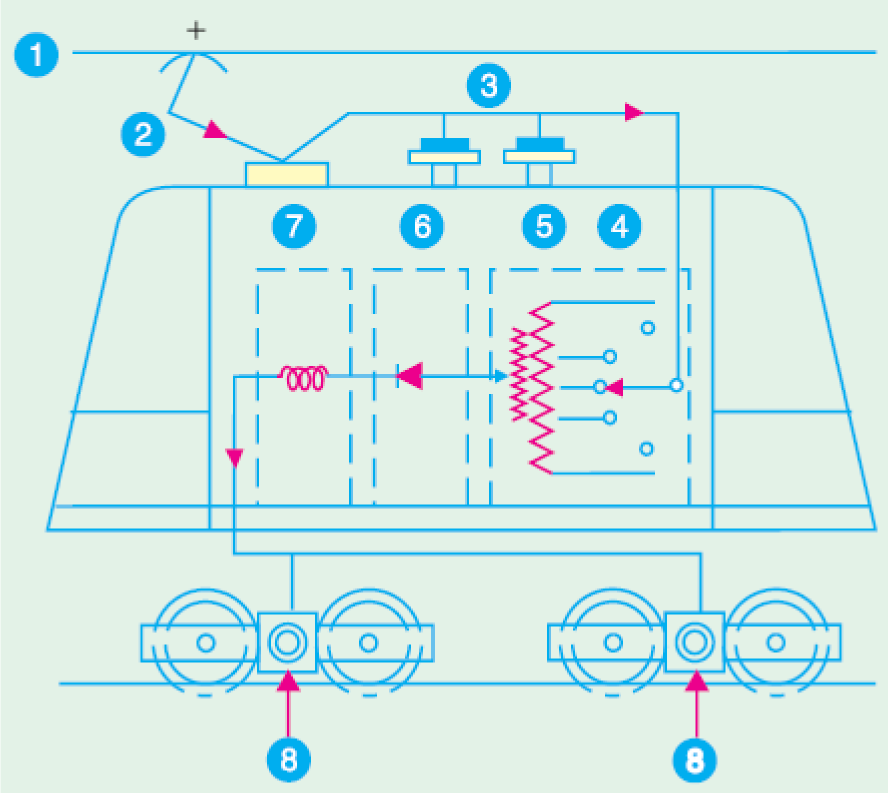


Fig. 2.1

The Tramways

It is the most economical means of transport for very dense traffic in the congested streets of large cities. It receives power through a bow collector or a grooved wheel from an overhead conductor at about 600 V dc, the running rail forming the return conductor. It is provided with at least two driving axles in order to ***(i)*** secure necessary adhesion ***(ii)*** start it from either end and ***(iii)*** use two motors with series-parallel control. Two drum-type controllers, one at each end, are used for controlling the tramcar. Though these controllers are connected in parallel, they have suitable interlocking arrangement meant to prevent their being used simultaneously.

Tramcars are being replaced by trolley-buses and internal-combustion-engined omnibuses because of the following reasons:

1. Tramcars lack flexibility of operation in congested areas.
2. The track constitutes a source of danger to other road users.

The Trolleybus

It is an electrically-operated pneumatic-tyred vehicle which needs ***no track in the roadway***. It receives its power at 600 V dc from two overhead contact wires. Since adhesion between a rubber tyred wheel and ground is sufficiently high, only a single driving axle and, hence, a single motor is used. The trolleybus can manoeuvre through traffic a metre or two on each side of the centre line of the trolley wires.

Trolley Bus

Overhead Equipment (OHE)

Broadly speaking, there are two systems of current collection by a traction unit:
(*i*) third rail system and **(*ii*)** overhead wire system.

It has been found that current collection from overhead wire is far superior to that from the third rail. Moreover, insulation of third rail at high voltage becomes an impracticable proposition and endangers the safety of the working personnel.

The simplest type of OHE consists of a single contact wire of hard drawn copper or silico-bronze supported either by bracket or an overhead span. To facilitate connection to the supports, the wire is grooved as shown in figure 2.2. Because there is appreciable sag of the wire between supports, it limits the speed of the traction unit to about 30 km/h. Hence, single contact wire system is suitable for tramways and in complicated yards and terminal stations where speeds are low and simplicity of layout is desirable.

For collection of current by high-speed trains, the contact (or trolley) wire has to be kept level without any abrupt changes in its height between the supporting structures. It can be done by using the single catenary system which consists of one catenary or messenger wire of steel with high sag and the trolley (or contact) wire supported from messenger wire by means of droppers clipped to both wires as shown in figure 2.3.

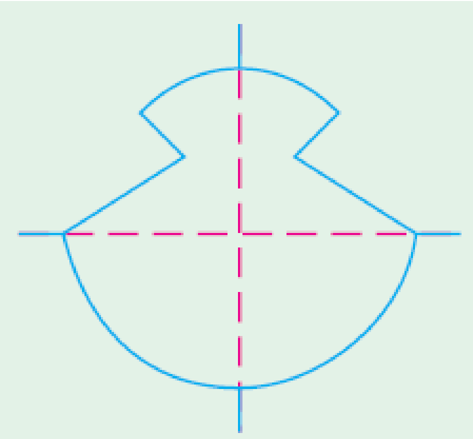


Fig. 2.2

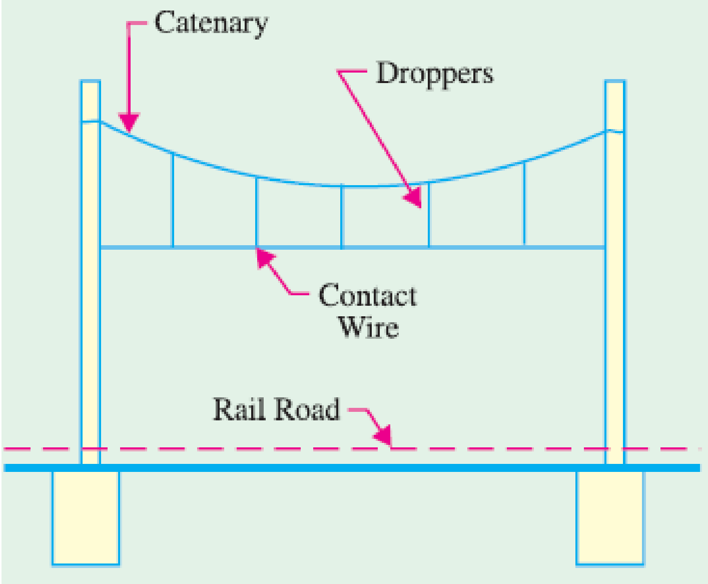


Fig. 2.3

Collector Gear for OHE

The most essential requirement of a collector is that it should keep continuous contact with trolley wire at all speeds. Three types of gear are in common use:

1. Trolley collector
2. Bow collector and
3. Pantograph collector.

To ensure even pressure on OHE, the gear equipment must, be flexible in order to follow variations in the sag of the contact wire. Also, reasonable precautions must be taken to prevent the collector from leaving the overhead wire at points and crossings.

The Trolley Collector

This collector is employed on tramways and trolley buses and is mounted on the roof of the vehicle. Contact with the OH wire is made by means of either a grooved wheel or a sliding shoe carried at the end of a light trolley pole attached to the top of the vehicle and held in contact with OH wire by means of a spring. The pole is hinged to a swivelling base so that it may be reversed for reverse running thereby making it unnecessary for the trolley wire to be accurately maintained above the centre of the track. Trolley collectors always operate in the trailing position. The trolley collector is suitable for low speeds upto 32 km/h beyond which there is a risk of its jumping off the OH contact wire particularly at points and crossing.

The Bow Collector

It can be used for higher speeds. As shown in figure 2.4, it consists of two roof mounted trolley poles at the ends of which is placed a light metal strip (or bow) about one metre long for current collection. The collection strip is purposely made of soft material (copper, aluminium or carbon) in order that most of the wear may occur on it rather than on the trolley wire. The bow collector also operates in the trailing position. Hence, it requires provision of either duplicate bows or an arrangement for reversing the bow for running in the reverse direction. Bow collector is not suitable for railway work where speeds up to 120 km/h and currents up to 3000 A are encountered. It is so because the inertia of the bow collector is too high to ensure satisfactory current collection.

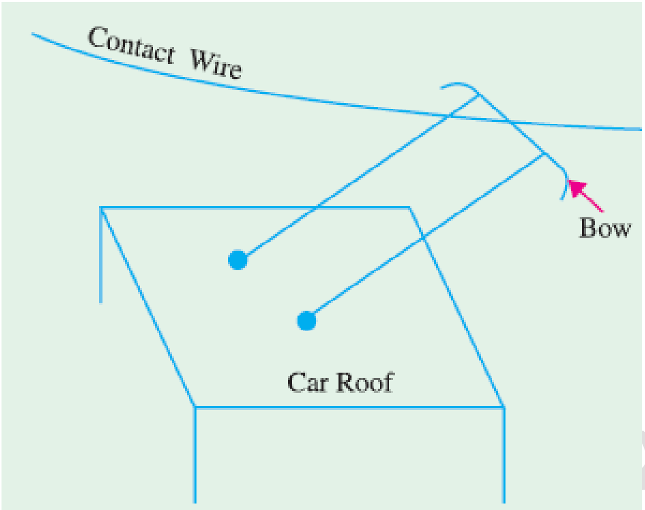


Fig. 2.4

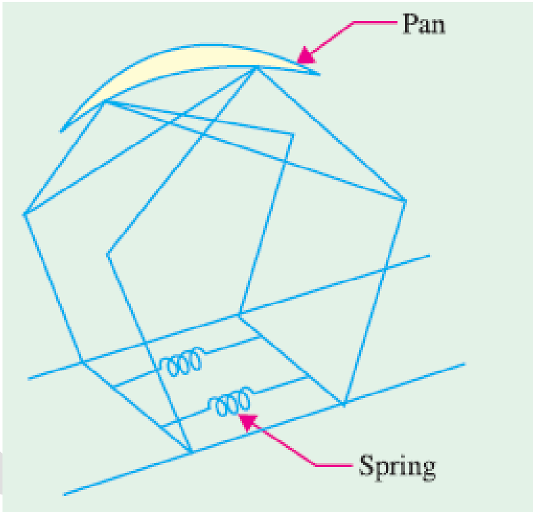


Fig. 2.5

The Pantograph Collector

Its function is to maintain link between overhead contact wire and power circuit of the electric locomotive at different speeds under all wind conditions and stiffness of OHE. It means that positive pressure has to be maintained at all times to avoid loss of contact and sparking but the pressure must be as low as possible in order to minimize wear of OH contact wire. A ‘diamond‘ type single-pan pantograph is shown in figure 2.5. It consists of a pentagonal framework of high-tensile alloy-steel tubing. The contact portion consists of a pressed steel pan fitted with renewable copper wearing strips which are forced against the OH contact wire by the upward action of pantograph springs. The pantograph can be raised or lowered from cabin by air cylinders.

Train Movement

The movement of trains and their energy consumption can be conveniently studied by means of speed/time and speed/distance curves. As their names indicate, former gives speed of the train at various ***times*** after the start of the run and the later gives speed at various ***distances*** from the starting point. Out of the two, speed/time curve is more important because

1. its slope gives acceleration or retardation as the case may be.
2. area between it and the horizontal (*i.e.* time) axis represents the distance travelled.
3. energy required for propulsion can be calculated if resistance to the motion of train is known.

Typical Speed/Time Curve

Typical speed/time curve for electric trains operating on passenger services is shown in figure 2.6. It may be divided into the following **five** parts :

1. Constant Acceleration Period (0 to t_1)

It is also called notching-up or starting period because during this period, starting resistance of the motors is gradually cut out so that the motor current (and hence, tractive effort) is maintained nearly constant which produces constant acceleration alternatively called ‘rheostatic acceleration‘ or ‘acceleration while notching‘.

2. Acceleration on Speed Curve (t_1 to t_2)

This acceleration commences after the starting resistance has been all cut-out at point *t*1 and full supply voltage has been applied to the motors. During this period, the motor current and torque decrease as train speed increases. Hence, acceleration gradually ***decreases*** till torque developed by motors exactly balances that due to resistance to the train motion. The shape of the portion *AB* of the speed/time curve depends primarily on the torque/speed characteristics of the traction motors.

3. Free-running Period (t_2 to t_3)

The train continues to run at the speed reached at point t_2. It is represented by portion *BC* in figure 2.6 and is a constant-speed period which occurs on level tracks.

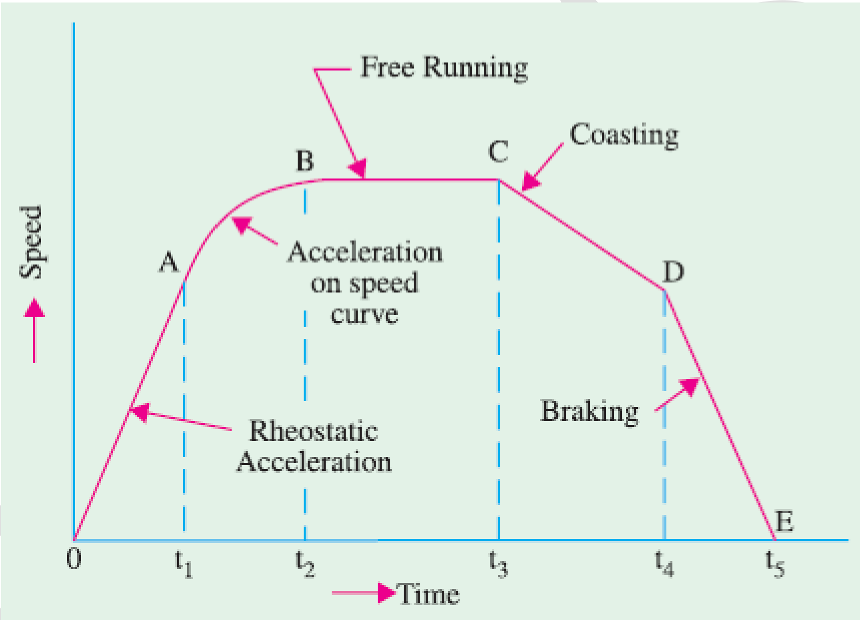


Fig. 2.6

4. Coasting (t_3 to t_4)

Power to the motors is cut off at point *t*3 so that the train runs under its momentum, the speed gradually falling due to friction, windage etc. (portion *CD*). During this period, retardation remains practically constant. Coasting is desirable because it utilizes some of the kinetic energy of the train which would, otherwise, be wasted during braking. Hence, it helps to reduce the energy consumption of the train.

5. Braking (t_4 to t_5)

At point t_4, brakes are applied and the train is brought to rest at point t_5.

It may be noted that coasting and braking are governed by train resistance and allowable retardation respectively.

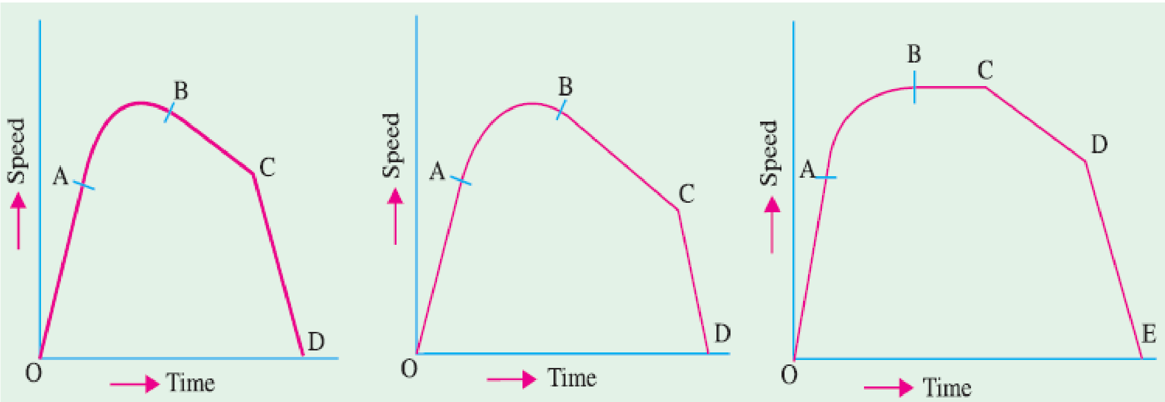
Speed/Time Curves for Different Services

Figure 2.7 (*a*) is representative of city service where relative values of acceleration and retardation are high in order to achieve moderately high average speed between stops.

Due to short Distances between stops, there is no possibility of free-running period though a short coasting period is included to save on energy consumption.

In suburban services [Fig. 2.7 (*b*)], again there is no free-running period but there is comparatively ***longer*** coasting period because of longer distances between stops. In this case also, relatively high values of acceleration and retardation are required in order to make the service as attractive as Possible.

For main-line service [Fig. 2.7 (c)], there are long periods of free-running at high speeds. The accelerating and retardation periods are relatively unimportant.



(a) (b) (c)

Fig. 2.7

Simplified Speed/Time Curve

For the purpose of comparative performance for a given service, the actual speed/time curve of figure 2.6 is replaced by a simplified speed/time curve which does not involve the knowledge of motor characteristics. Such a curve has simple geometric shape so that simple mathematics can be used to find the relation between acceleration, retardation, average speed and distance etc. The simple curve would be fairly accurate provided it **(*i*)** ***retains the same acceleration and retardation and (ii) has the same area as the actual speed/time curve***. The simplified speed/time curve can have either of the two shapes:

(i) trapezoidal shape OA_1B_1C of figure 2.8 where speed-curve running and coasting periods of the actual speed/time curve have been replaced by a constant speed period.

(ii) Quadrilateral shape OA_2B_2C where the same two periods are replaced by the extensions of initial constant acceleration and coasting periods. It is found that trapezoidal diagram OA_1B_1C gives simpler relationships between the principal quantities involved in train movement and also gives closer approximation of actual energy consumed during ***main-line service on level track***. On the other hand, quadrilateral diagram approximates more closely to the actual conditions in ***city and suburban services***.

Average and Schedule Speed

While considering train movement, the following three speeds are of importance:

1. Crest Speed. It is the maximum speed (V_m) attained by a train during the run.

2. Average Speed = $\dfrac{\text{distance between stops}}{\text{actual time of run}}$

In this case, only running time is considered but *not the stop time*.

3. Schedule Speed = $\dfrac{\text{distance between stops}}{\text{actual time of run} + \text{stop time}}$

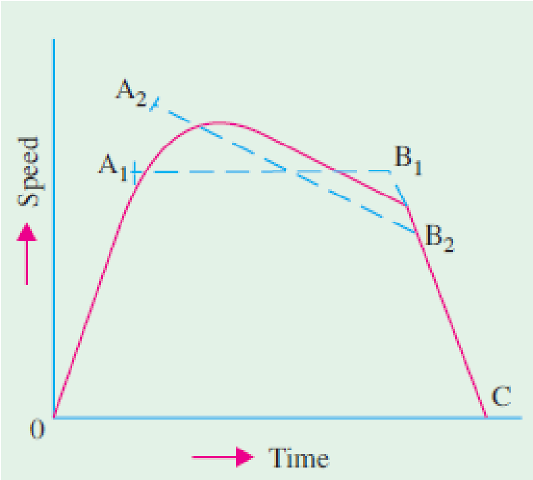


Fig. 2.8

Obviously, schedule speed can be obtained from average speed by including the duration of stops. For a given distance between stations, higher values of acceleration and retardation will mean lesser running time and, consequently, higher schedule speed. Similarly, for a given distance between stations and for fixed values of acceleration and retardation, higher crest speed will result in higher schedule speed. For the same value of average speed, increase in duration of stops decreases the schedule speed.

SI Units in Traction Mechanics

In describing various quantities involved in the mechanics of train movement, only the latest SI system will be used. Since SI system is an 'absolute system', only absolute units will be used while gravitational units (used hitherto) will be discarded.

1. **Force.** It is measured in newton (N)
2. **Mass.** Its unit is kilogram (kg). Commonly used bigger units is tonne (*t*),
   1 tonne = 1000kg
3. **Energy.** Its basic unit is joule (J). Other units often employed are watt-hour (Wh) and kilowatthour (kWh).
   $$1\ \text{Wh} = 1\frac{\text{J}}{\text{s}} \times 3600\ \text{s} = 3600\ \text{J} = 3.6\ \text{kJ}$$
   $$1\ \text{kWh} = 1000 \times 1\frac{\text{J}}{\text{s}} \times 3600\ \text{s} = 36 \times 10^5\ \text{J} = 3.6\ \text{MJ}$$
4. **Work.** Its unit is the same as that of energy.
5. **Power.** Its unit is watt (W) which equals 1 J/s. Other units are kilowatt (kW) and megawatt (MW).
6. **Distance.** Its unit is metre. Other unit often used is kilometre (km).
7. **Velocity.** Its absolute unit is metre per second (m/s). If velocity is given in km/h (or km.ph), it can be easily converted into the SI unit of m/s by multiplying it with a factor of (1000/3600) = 5/18 = 0.2778.
   For example, 72 km.ph = 72 × 5/18 = 72 × 0.2778 = 20 m/s.
8. **Acceleration.** Its unit is metre/second^2 (m/s^2). If acceleration is given in km/h/s (or

km.ph.ps), then it can be converted into m/s^2 by simply multiplying it by the factor (1000/3600) = 5/18 = 0.2778 *i.e.* the same factor as for velocity.

For example, 1.8 km.ph.ps = 1.8 × 5/18 = 1.8 × 0.2778 = 0.5 m/s^2

Confusion Regarding Weight and Mass of a Train

Many students often get confused regarding the correct meaning of the terms 'weight' and 'mass' and their units while solving numericals on train movement particularly when they are not expressed clearly and consistently in their absolute units. It is primarily due to the mixing up of absolute units with gravitational units. There would be no confusion at all if ***we are consistent in using only absolute units*** as required by the SI system of units which disallows the use of gravitational units.

Though this topic was briefly discussed earlier, it is worth repeating here.

1. **Mass (M).** It is the quantity of matter contained in a body.
   Its absolute unit is kilogram (kg). Other multiple in common use is tonne.
2. **Weight (W).** It is the **force** with which earth pulls a body downwards.

The weight of a body can be expressed in (*i*) the ***absolute*** unit of newton (N) or (*ii*) the ***gravitational*** unit of kilogram-weight (kg. wt) which is often writing as 'kgf' in engineering literature.

Another still bigger ***gravitational*** unit commonly used in traction work is tonne-weight (t-wt)

$$1 \text{ t-wt} = 1000 \text{ kg-wt} = 1000 \times 9.8 \text{ N} = 9800 \text{ N}$$

(*i*) Absolute Unit of Weight

It is called newton (N) whose definition may be obtained from Newton's Second Law of Motion.

Commonly used multiple is kilo-newton (kN). Obviously, 1 kN = 1000 N = 103 N.

For example, if a mass of 200 kg has to be given an acceleration of 2.5 m/s^2, force required is F = 200 × 2.5 = 500 N.

If a train of mass 500 tonne has to be given an acceleration of 0.6 m/s2, force required is

$$F = ma = (500 \times 1000) \times 0.6 = 300{,}000 \text{ N} = 300 \text{ kN}$$

(*ii*) Gravitational Unit of Weight

It is 'g' times bigger than newton. It is called kilogram-weight (kg.wt.)

$$1 \text{ kg.wt} = g \text{ newton} = 9.81 \text{ N \_ } 9.8 \text{ N}$$

Unfortunately, the word 'wt' is usually omitted from kg-wt when expressing the weight of the body on the assumption that it can be understood or inferred from the language used.

Take the statement "a body has a *weight* of 100 kg". It looks as if the weight of the body has been expressed in terms of the mass unit 'kg'. To avoid this confusion, statement should be 'a body has a weight of 100 kg. wt.' But the first statement is justified by the writers on the ground that since the word 'weight' has already been used in the statement, it should be automatically understood by the readers that 'kg' is not the 'kg' of mass but is kg-wt. It would be mass kg if the statement is 'a body has a mass of 100 kg'. Often kg-wt is written as 'kgf' where 'f' is the first letter of the word force and is added to distinguish it from kg of mass.

Now, consider the statement "a body *weighing* 500 kg travels with a speed of 36 km/h.........."

Now, weight of the body W = 500 kg.wt. = 500 × 9.8 N

Since we know the weight of the body, we can find its mass from the relation $W = mg$. But while using this equation, it is essential that we must ***consistently*** **use** the ***absolute*** units only. In this equation, W must be in newton (not in kg. wt), m in kg and g in m/s^2.

$\therefore$ 500 × 9.8 = m × 9.8 ; $\therefore$ m = 500 kg

It means that a body which *weighs* 500 kg (wt) has a *mass* of 500 kg.

As a practical rule, weight of a body in **gravitational** units is numerically equal to its mass in **absolute** units. This simple fact must be clearly understood to avoid any confusion between weight and mass of a body.

A train which weighs 500 tonne has a mass of 500 tonne as proved below :

train weight, W = 500 tonne-wt = 500 × 1000 kg-wt = 500 × 1000 × 9.8 N

Now, $W = mg$; $\therefore$ 500 × 1000 × 9.8 = m × 9.8

$\therefore$ m = 500 × 1000 kg = 500 × 1000/1000 = 500 tonne

To avoid this unfortunate confusion, it would be helpful to change our terminology. For example, instead of saying "a train weighing 500 tonne is........" it is better to say "a 500-t train is ...." or "a train having a mass of 500 t is ...."

In order to remove this confusion, SI system of units has disallowed the use of gravitational units. There will be no confusion if ***we consistently use only absolute units***.

Quantities Involved in Traction Mechanics

Following principal quantities are involved in train movement:

D = distance between stops

M = dead mass of the train

M_e = effective mass of the train

W = dead weight of the train

W_e = effective weight of the train

α = acceleration during starting period

β_c = retardation during coasting

β = retardation during braking

V_a = average speed

V_m = maximum (or crest) speed.

t = total time for the run

t_1 = time of acceleration

t_2 = time of free running = $t - (t_1 + t_3)$

t_3 = time of braking

F_t = tractive effort T = torque

Relationship between Principal Quantities in Trapezoidal Diagram

As seen from figure 2.9.

$\alpha = V_m / t_1$ or $t_1 = V_m / \alpha$

$\beta = V_m / t_3$ or $t_3 = V_m / \beta$

As we know, total distance D between the two stops is given by the area of trapezium $OABC$.

$\therefore$ D = area $OABC$

= area OAD + area $ABED$ + area BCE

$$= \frac{1}{2} V_m t_1 + V_m t_2 + \frac{1}{2} V_m t_3$$

$$= \frac{1}{2} V_m t_1 + V_m [t - (t_1 + t_3)] + \frac{1}{2} V_m t_3$$

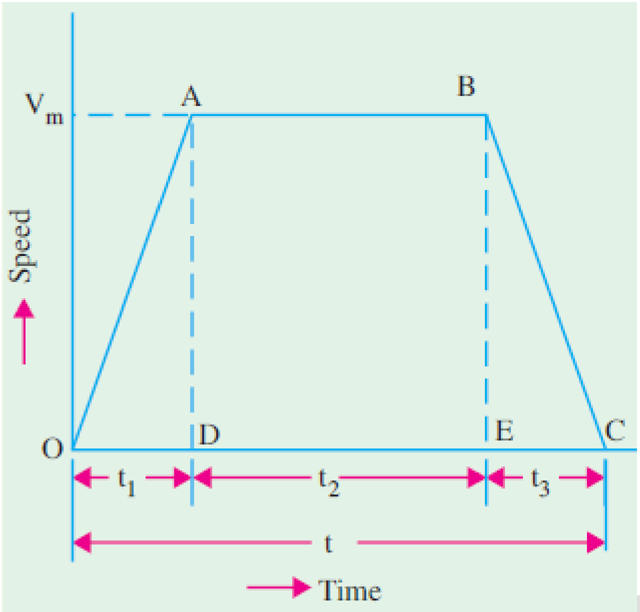


Fig. 2.9

$$D = V_m\left[\frac{t_1}{2} + t - t_1 - t_3 + \frac{t_3}{2}\right]$$

$$= V_m\left[t - \frac{1}{2}(t_1 + t_3)\right]$$

$$= V_m\left[t - \frac{V_m}{2}\left(\frac{1}{\alpha} + \frac{1}{\beta}\right)\right]$$

$$Let, \quad K = \frac{1}{2}\left(\frac{1}{\alpha} + \frac{1}{\beta}\right)$$

Substituting this value of K in the above equation, we get
$D = V_m(t - KV_m)$

or $KV_m^2 - V_m t + D = 0$ ... ... (i)

$$\therefore \quad V_m = \frac{t \pm \sqrt{t^2 - 4KD}}{2K}$$

Rejecting the positive sign which gives impracticable value, we get

$$V_m = \frac{t \pm \sqrt{t^2 - 4KD}}{2K}$$

From equation (i) above, we get

$$KV_m^2 = V_m t - D \qquad \text{or} \qquad K = \frac{t}{V_m} - \frac{D}{V_m^2} = \frac{D}{V_m^2}\left(V_m.\frac{t}{D} - 1\right)$$

Now,

$$\boldsymbol{V_a = \frac{D}{t}}$$

$$\therefore \quad \boldsymbol{K = \frac{D}{V_m^2}\left(\frac{V_m}{V_a} - 1\right)}$$

Obviously, if V_m, V_a and D are given, then value of K and hence of α and β can be found.

Relationship between Principal Quantities in Quadrilateral Diagram

The diagram is shown in figure2.10. Let β_c represent the retardation during coasting period. As before,

$$t_1 = V_1/\alpha,\ t_2 = (V_1 - V_2)/\beta_c \text{ and } t_3 = V_2/\beta$$

$$\begin{aligned}
D &= \text{area } OABC \\
&= \text{area } OAD + \text{area } ABED + \text{area } BCE \\
&= \frac{1}{2}V_1 t_1 + t_2\left(\frac{V_1+V_2}{2}\right) + \frac{1}{2}V_2 t_3 \\
&= \frac{1}{2}V_1(t_1+t_2) + \frac{1}{2}V_2(t_2+t_3) \\
&= \frac{1}{2}V_1(t-t_3) + \frac{1}{2}V_2(t-t_1) \\
&= \frac{1}{2}t(V_1+V_2) - \frac{V_2 t_1}{2} - \frac{V_1 t_3}{2} \\
&= \frac{1}{2}t(V_1+V_2) - \frac{1}{2}V_1V_2\left(\frac{1}{\alpha}+\frac{1}{\beta}\right) \\
&= \frac{1}{2}t(V_1+V_2) - KV_1V_2
\end{aligned}$$

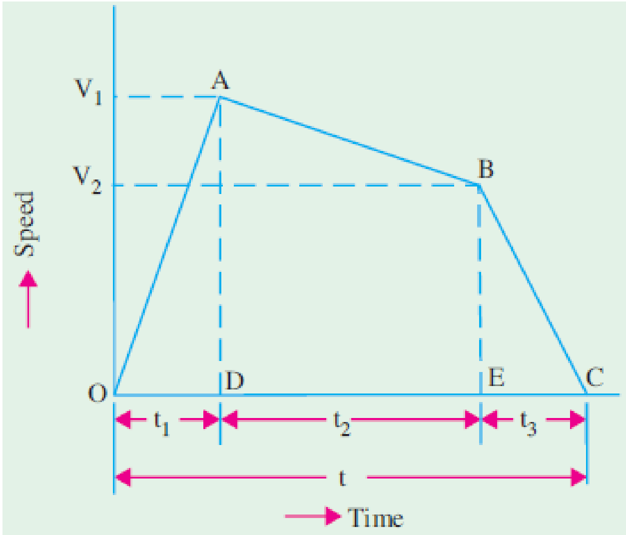


Fig. 2.10

where $K = \frac{1}{2}\left(\frac{1}{\alpha}+\frac{1}{\beta}\right) = \frac{\alpha+\beta}{2\alpha\beta}$ Also, $\beta_c = \frac{(V_1 - V_2)}{t_2}$

$$\therefore \quad V_2 = V_1 - \beta_c t_2 = V_1 - \beta_c(t - t_1 - t_3)$$

$$= V_1 - \beta_c\left(t - \frac{V_1}{\alpha} - \frac{V_2}{\beta}\right) = V_1\beta_c\left(t - \frac{V_1}{\alpha}\right) + \beta_c\frac{V_2}{\beta}$$

or $V_2\left(1 - \frac{\beta_c}{\beta}\right) = V_1 - \beta_c\left(t - \frac{V_1}{\alpha}\right) \qquad \therefore \quad V_2 = \frac{V_1 - \beta_c(t - V_1/\alpha)}{(1 - \beta_c/\beta)}$

Example 2.1. A suburban train runs with an average speed of 36 km/h between two stations 2 km apart. Values of acceleration and retardation are 1.8 km/h/s and 3.6 km/h/s. compute the maximum speed of the train assuming trapezoidal speed/time curve.

Solution. Now, $V_a = 36$ km/h $= 36 \times 5/18 = 10$ m/s

$$\alpha = 1.8 \text{ km/h/s} = 1.8 \times 5/18 = 0.5 \text{ m/s}^2,\ \beta = 3.6 \text{ km/h/s} = 3.6 \times 5/18 = 1.0 \text{ m/s}^2$$

$$t = D/V_a = 2000/10 = 200 \text{ s};\ K = (\alpha+\beta)/2\alpha\beta = (0.5+1.0)/2 \times 0.5 \times 1 = 1.5$$

$$V_m = \frac{t - \sqrt{t^2 - 4KD}}{2K} = \frac{200 - \sqrt{200^2 - 4 \times 1.5 \times 2000}}{2 \times 1.5}$$

$$= 11 \text{ m/s} = 11 \times 18/5$$

$$= \mathbf{39.6\ km/h}$$

Example 2.2. A train is required to run between two stations 1.5 km apart at a schedule speed of 36 km/h, the duration of stops being 25 seconds. The braking retardation is 3 km/h/s. Assuming a trapezoidal speed/time curve, calculate the acceleration if the ratio of maximum speed to average speed is to be 1.25.

Solution. Here, $D = 1500$ m ; schedule speed = 36 km/h = 36 × 5/18 = 10 m/s

$$\beta = 3 \text{ km/h/s} = 3 \times 5/18 = 5/6 \text{ m/s}^2$$

Schedule time of run $= 1500/10 = 150$ s ; Actual time of run $= 150 - 25 = 125$ s

$\therefore \quad V_a = 1500/125 = 12$ m/s ; $V_m = 1.25 \times 12 = 15$ m/s

Now, $$K = \frac{D}{V_m^2}\left(\frac{V_m}{V_a} - 1\right) = \frac{1500}{15^2}(1.25 - 1) = \frac{5}{3}$$

Also, $$K = \frac{1}{2}\left(\frac{1}{\alpha} + \frac{1}{\beta}\right) \quad \text{or} \quad \frac{5}{3} = \frac{1}{2}\left(\frac{1}{\alpha} + \frac{6}{5}\right)$$

$\therefore \quad \alpha = 0.47 \text{ m/s}^2 = 0.47 \times 18/5 =$ **1.7 km/h/s**

Example 2.3. Find the schedule speed of an electric train for a run of 1.5 km if the ratio of its maximum to average speed is 1.25. It has a braking retardation of 3.6 km/h/s, acceleration of 1.8 km/h/s and stop time of 21 second. Assume trapezoidal speed/time curve.

Solution. $\alpha = 1.8 \times 5/18 = 0.5 \text{ m/s}^2, \beta = 3.6 \times 5/18 = 1.0 \text{ m/s}^2$

$D = 1.5 \text{ km} = 1500 \text{ m}$

$$K = \frac{1}{2}\left(\frac{1}{0.5} + \frac{1}{1}\right) = \frac{3}{2} \qquad \text{Now,} \quad K = \frac{D}{V_m^2}\left(\frac{V_m}{V_a} - 1\right)$$

or $$V_m^2 = \frac{D}{K}\left(\frac{V_m}{V_a} - 1\right) \qquad \therefore \quad V_m^2 = \frac{1500}{3/2}(1.25 - 1) = 250 \text{ ; } V_m = 15.8 \text{ m/s}$$

$V_a = V_m/1.25 = 15.8/1.25 = 12.6$ m/s

Actual time of run $= 1500/12.6 = 119$ seconds

Schedule time $= 119 + 21 = 140$ second

$\therefore$ Schedule speed $= 1500/140 = 10.7$ m/s $=$ **38.5 km/h**

Example 2.4. A train runs between two stations 1.6 km apart at an average speed of 36 km/h. If the maximum speed is to be limited to 72 km/h, acceleration to 2.7 km/h/s, coasting retardation to 0.18 km/h/s and braking retardation to 3.2 km/h/s, compute the duration of acceleration, coasting and braking periods.

Assume a simplified speed/time curve.

Solution. Given : $D = 1.6 \text{ km} = 1600 \text{ m}, \quad V_a = 36 \text{ km/h} = 10 \text{ m/s}$

$V_1 = 72 \text{ km/h} = 20 \text{ m/s ;} \quad \alpha = 2.7 \text{ km/h/s} = 0.75 \text{ m/s}^2$

$\beta_c = 0.18 \text{ km/h/s} = 0.05 \text{ m/s}^2 \text{ ;} \quad \beta = 3.6 \text{ km/h/s} = 1.0 \text{ m/s}^2$

With reference to Fig. 43.12, we have

Duration of acceleration, $t_1 = V_1/\alpha = 20/0.75 = 27$ s

Actual time of run, $t = 1600/10 =$ **160 s**

Duration of braking, $t_3 = V_2/1.0 = V_2$ second

Duration of coasting, $t_2 = (V_1 - V_2)/\beta_c = (20 - V_2)/ 0.05 = (400 - 20 V_2)$ second

Now, $t = t_1 + t_2 + t_3$ or $160 = 27 + (400 - 20 V_2) + V_2 \quad \therefore \quad V_2 = 14$ m/s

$\therefore \quad t_2 = (20 - 14)/0.05 = 120$ s ; $t_3 = 14/1.0 =$ **14 s**

Tractive Effort for Propulsion of a Train:

The tractive effort (F_t) is the force developed by the traction unit at the rim of the driving wheels for moving the unit itself and its train (trailing load). The tractive effort required for train propulsion on a *level track* is

$F_t = F_a + F_r$

If gradients are involved, the above expression becomes

$F_t = F_a + F_g + F_r$ — for ascending gradient

$= F_a - F_g + F_r$ — for descending gradient

where F_a = force required for giving linear acceleration to the train

F_g = force required to overcome the effect of gravity

F_r = force required to overcome resistance to train motion.

(*a*) Value of F_a

If M is the dead (or stationary) mass of the train and a its linear acceleration, then

$F_a = Ma$

Since a train has rotating parts like wheels, axles, motor armatures and gearing etc., its *effective* (or accelerating) mass Me is more (about 8 − 15%) than its stationary mass. These parts have to be given angular acceleration at the same time as the whole train is accelerated in the linear direction.

Hence, $F_e = M_e a$

(*i*) If M_e is in kg and α in m/s^2, then $Fa = Me\ a$ newton

(*ii*) If M_e is in tonne and α in km/h/s, then converting them into absolute units, we have

$F_a = (1000\ M_e) \times (1000/3600)\ a = 277.8\ M_e\ a$ newton

(*b*) Value of F_g

As seen from figure 2.11, $F_g = W \sin\theta = Mg \sin\theta$

In railway practice, gradient is expressed as the rise (in metres) a track distance of 100 m and is called percentage gradient.

$$\therefore \qquad \%\ G = \frac{BC}{AC/100} = 100\frac{BC}{AC} = 100 \sin\theta$$

Substituting the value of sinθ in the above equation, we get

$$F_g = Mg\ G/100 = 9.8 \times 10^{-2}\ MG$$

(*i*) When M is in kg, $F_g = 9.8 \times 10^{-2}\ MG$ newton

(*ii*) When M is given in tonne, then

$$F_g = 9.8 \times 10^{-2}\ (1000\ M)\ G = 98\ MG \text{ newton}$$

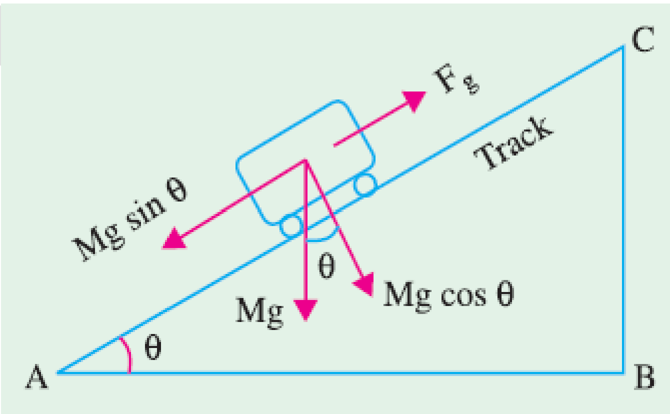


Fig. 2.11

(*c*) Value of Fr

Train resistance comprises all those forces which oppose its motion. It consists of mechanical resistance and wind resistance. Mechanical resistance itself is made up of internal and external resistances. The internal resistance comprises friction at journals, axles, guides and buffers etc. The external resistance consists of friction between wheels and rails and flange friction etc. Mechanical resistance is almost independent of train speed but depends on its weight. The wind friction varies directly as the square of the train speed.

If r is specific resistance of the train *i.e.* resistance offered per unit mass of the train, then $F_r = M.r.$

(i) If r is in newton per kg of train mass and M is the train mass in kg, then

$$F_r = M.r \text{ newton}$$

(ii) If r is in newton per tonne train mass (N/t) and M is in tonne (t), then

$$F_r = M \text{ tonne} \times r = M_r \text{ newton*}$$

Hence, expression for total tractive effort becomes

$$F_t = F_a \pm F_g + F_r = (277.8\ \alpha\ Me \pm 98\ MG + Mr) \text{ newton}$$

Please remember that here M is in tonne, α in km/h/s, G is in metres per 100 m of track length (*i.e.* % G) and r is in newton/tonne (N/t) of train mass.

The positive sign for F_g is taken when motion is along an ascending gradient and negative sign when motion is along a descending gradient.

* If r is in kg (wt) per tonne train mass and M is in tonne, then $F_r = M$ tonne $\times$ ($r \times 9.8$) newton/tonne = 9.8 Mr newton.

Power Output from Driving Axles:

If F_t is the tractive effort and ν is the train velocity, then

$$\text{output power} = F_t \times \nu$$

(*i*) If F_t is in newton and ν in m/s, then

$$\text{output power} = F_t \times \nu \text{ watt}$$

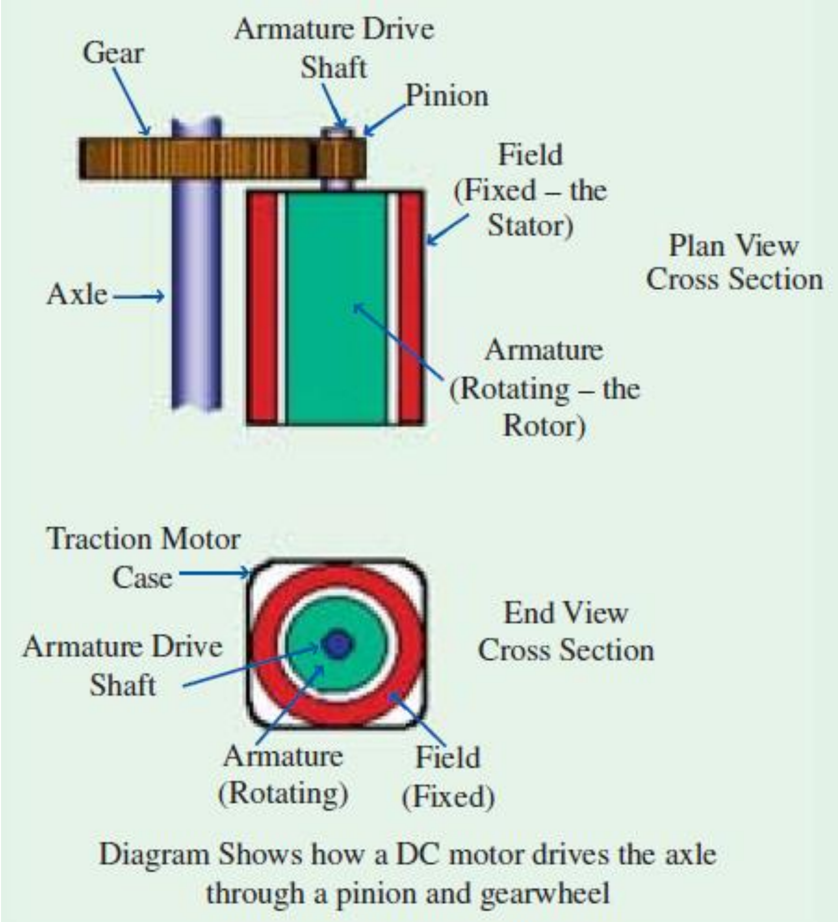
(*ii*) If F_t is in newton and ν is in km/h , then converting ν into m/s, we have

$$\text{output power} = F_t \times \left(\frac{1000}{3600}\right)\nu \text{ watt} = \frac{F_t\,\nu}{3600} \text{ kW}$$

If η is the efficiency of transmission gear, then power output of motors is

$$= F_t\,.\,\nu/\eta \text{ watt} \qquad \text{— } \nu \text{ in m/s}$$

$$= \frac{F_t\,\nu}{3600\,\eta} \text{ kW} \qquad \text{— } \nu \text{ in km/h}$$



Traction motor schematic diagram

Energy Output from Driving Axles:

Energy (like work) is given by the product of power and time.

$$E = (F_t \times v) \times t = F_t \times (v \times t) = F_t \times D$$

where D is the distance travelled in the direction of tractive effort.

Total energy output from driving axles for the run is

E = Energy during acceleration + Energy during free run

As seen from figure 2.12

$$E = F_t \times area\ OAD + F_t' \times area\ ABED$$

$$E = F_t \times \frac{1}{2} V_m t_1 + F'_t \times V_m t_2$$

where F_t is the tractive effort during accelerating period and F_t' that during free running period. Incidentally, F_t will consist of all the three components (F_a, F_g, F_r **),** whereas F_t' will consist of (98 MG + Mr) provided there is an ascending gradient.

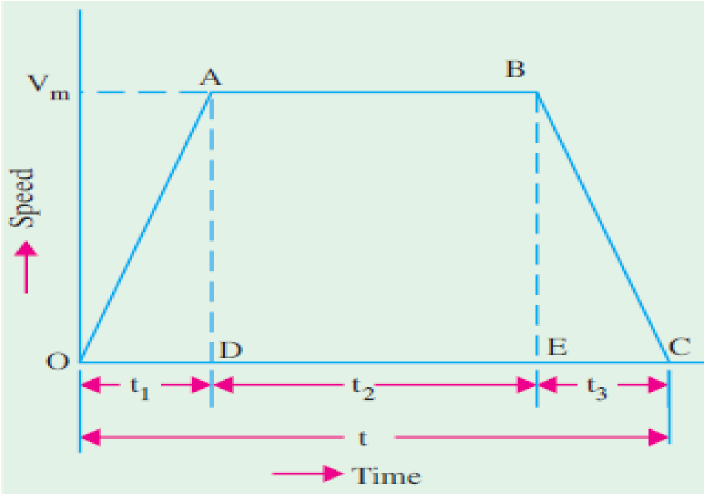


Fig. 2.12

Specific Energy Output:

It is the energy output of the driving wheel expressed in watt-hour (Wh) per tonne-km (*t*-km) of the train. It can be found by first converting the energy output into Wh and then dividing it by the mass of the train in tonne and route distance in km.

Hence, unit of specific energy output generally used in railway work is : Wh/tonne-km (Wh/t-km).

Evaluation of Specific Energy Output

We will first calculate the total energy output of the driving axles and then divide it by train mass in tonne and route length in km to find the specific energy output. It will be presumed that :

(i) there is a gradient of G throughout the run and

(ii) power remains ON upto the end of free run in the case of trapezoidal curve and upto the accelerating period in the case of quadrilateral curve

Now, output of the driving axles is used for the following purposes:

1. for accelerating the train

2. for overcoming the gradient

3. for overcoming train resistance.

(*a*) Energy required for train acceleration (*Ea*)

As seen from trapezoidal diagram (Fig. 2.12)

$$E_a = F_a \times \text{distance } OAD = 277.8\,\alpha\, M_e \times \frac{1}{2} V_m . t_1 \text{ joules}$$

$$= 277.8\,\alpha\, M_e \times \frac{1}{2} V_m \times \frac{V_m}{\alpha} \text{ joules} \qquad \left(\because t_1 = \frac{V_m}{\alpha} \right)$$

$$= 277.8\,\alpha\, M_e \times \left[\frac{1}{2} . \frac{V_m \times 1000}{3600} \times \frac{V_m}{\alpha} \right] \text{ joules}$$

It will be seen that since V_m is in km/h, it has been converted into m/s by multiplying it with the conversion factor of (1000/3600). In the case of (V_m /*t*), conversion factors for V_m and α being the same, they cancel out. Since 1 Wh = 3600 J.

$$E_a = 277.8\,\alpha\,M_e \left[\frac{1}{2}.\frac{V_m \times 1000}{3600} \times \frac{V_m}{\alpha}\right] \text{Wh} = 0.01072\,\frac{V_m^2}{M_e}\ \text{Wh}$$

(b) Energy required for overcoming gradient (E_g)

$$E_g = F_g \times D'$$

where ***D'*** is the ***total distance over which power remains*** ON. Its maximum value equals the distance represented by the area *OABE. i.e.* from the start to the end of free-running period in the case of trapezoidal curve [as per assumption (*i*)].

Substituting the value of F_g, we get

$$E_g = 98\ MG.\ (1000\ D)\ \text{joules} = 98{,}000\ MGD\ \text{joules}$$

It has been assumed that ***D*** is in km.

When expressed in Wh, it becomes

$$E_g = 98{,}000\ MGD' \frac{1}{3600}\ \text{Wh} = 27.25\ MGD'\ \text{Wh}$$

(c) Energy required for overcoming resistance (E_r)

$$E_r = F_r \times D' = M\,.\,r \times (1000\ D')\ \text{joules} \qquad \text{— } D' \text{ in km}$$

$$= \frac{1000\ Mr\ D'}{3600}\ \text{Wh} = 0.2778\ Mr\ D'\ \text{Wh} \qquad \text{— } D' \text{ in km}$$

∴ total energy output of the driving axles is

$$E = E_a + E_g + E_r$$

$$= (0.01072\ V_m^2/M_e + 27.25\ MGD' + 0.2778\ Mr\ D'\ \text{Wh}$$

Specific energy output

$$E_{spo} = \frac{E}{M \times D} \qquad \text{— } D \text{ is the } \textit{total} \text{ run length}$$

$$= \left(0.01072\,\frac{V_m^2}{D}.\frac{M_e}{M} + 27.25\ G\,\frac{D'}{D} + 0.2778\ r\,\frac{D'}{D}\right) \text{Wh/t-km}$$

It may be noted that *if there is no gradient*, then

$$E_{spo} = \left(0.01072\,\frac{V_m^2}{D}.\frac{M_e}{M} + 0.2778\ r\,\frac{D'}{D}\right) \text{Wh/t-km}$$

Alternative Method:

As before, we will consider the trapezoidal speed/time curve. Now, we will calculate energy output not **force-wise** but **period-wise**.

(i) Energy output during accelerating period

$$
\begin{aligned}
E_a &= F_t \times \text{distance travelled during accelerating period} \\
&= F_t \times \text{area } OAD \\
&= F_t \times \frac{1}{2} V_m\, t_1 = \frac{1}{2} F_t \,.V_m \,.\frac{V_m}{\alpha} \\
&= \frac{1}{2}.F_t \left(\frac{1000}{3600}.V_m\right).\frac{V_m}{\alpha} \text{ joules} \\
&= \frac{1}{2}.F_t \left(\frac{1000}{3600}.V_m\right).\frac{V_m}{\alpha}.\frac{1}{3600} \text{ Wh}
\end{aligned}
$$

Substituting the value of F_t , we get

$$
E_a = \frac{1000}{(3600)^2} \cdot \frac{V_m^2}{2\alpha}(277.8\,\alpha\, M_e + 98\, MG + Mr) \text{ Wh}
$$

It must be remembered that during this period, ***all the three forces are at work.***

(ii) Energy output during free-running period

Here, work is required only against two forces *i.e.* gravity and resistance (as mentioned earlier).

$$
\begin{aligned}
\text{Energy} \quad E_{fr} &= F_t' \times \text{area } ABED \\
&= F_t' \times (V_m \times t_2) = F_t' \times \left(\frac{1000}{3600} V_m\right) . t_2 \text{ joules} \\
&= F_t' \times \left(\frac{1000}{3600} V_m\right) \times t_2 \times \frac{1}{3600} \text{ Wh} = \left(\frac{1000}{3600}\right) F_t' \times V_m\, t_2 \,.\frac{1}{3600} \text{ Wh} \\
&= \left(\frac{1000}{3600}\right). F_t' \times D_{fr} \text{ Wh} = \left(\frac{1000}{3600}\right)(98\, MG + Mr)\, D_{fr} \text{ Wh}
\end{aligned}
$$

where D_{fr} is the distance in km travelled during the free-running period*

Total energy required is the sum of the above two energies.

$$
\begin{aligned}
\therefore \quad E &= E_a + E_{fr} \\
&= \frac{1000}{(3600)^2} \frac{V_m^2}{2\,\alpha}(277.8\,\alpha\,.\,M_e + 98\,MG + Mr) + \frac{1000}{3600}(98\,MG + Mr)\,D_{fr} \text{ Wh} \\
&= \frac{1000}{(3600)^2} \frac{V_m^2}{2\,\alpha}\, 277.8\,\alpha\, M_e + \frac{1000}{(3600)^2}.\frac{V_m^2}{2\,\alpha}(98\,MG + Mr) + \frac{1000}{3600}(98\,MG + Mr)\,.\,D_{fr} \text{ Wh} \\
&= 0.01072\, V_m^2\,.\,M_e + \frac{1000}{3600}(98\,MG + Mr)\left(\frac{V_m^2}{2\alpha \times 3600} + D_{fr}\right) \text{Wh}
\end{aligned}
$$

$$
\text{Now,} \quad \frac{V_m^2}{2\alpha \times 3600} = \frac{1}{2}\left(\frac{V_m}{3600}\right).\frac{V_m}{\alpha} = \frac{1}{2}\left(\frac{V_m}{3600}\right).t_1
$$

$$
= \text{distance travelled during accelerating period } i.e.\ D_a
$$

$$
\therefore \quad E = 0.01072\, V_m^2\,.\,M_e + \frac{1000}{3600}(98\,MG + Mr)(D_a + D_{fr}) \text{ Wh}
$$

$$
= 0.01072\, V_m^2\,.\,M_e + (27.25\,MG + 0.2778\,Mr)\,D' \text{ Wh}
$$

It is the same expression as found above.

---

* D_{fr} = velocity in km/h × time in hours

= $V_m \times (t_2/3600)$ because times are always taken in seconds.

Energy Consumption:

It equals the total energy input to the traction motors from the supply. It is usually expressed in Wh which equals 3600 J. It can be found by dividing the energy output of the driving wheels with the combined efficiency of transmission gear and motor.

$$\therefore \; Energy\ consumption = \frac{Output\ of\ driving\ axles}{\eta_{motor} \times \eta_{gear}}$$

Specific Energy Consumption:

It is the energy consumed (in Wh) per tonne mass of the train per km length of the run.

Specific energy consumption,

$$E_{spc} = \frac{\text{total energy consumed in Wh}}{\text{train mass in tonne} \times \text{run length in km}} = \frac{\text{specific energy output}}{\eta}$$

where η = overall efficiency of transmission gear and motor $= \eta_{gear} \times \eta_{motor}$

$$E_{spc} = \left(0.01072 \cdot \frac{V_m^2}{\eta D} \cdot \frac{M_e}{M} + 27.25 \frac{G}{\eta} \cdot \frac{D'}{D} + 0.2778 \frac{r}{\eta} \cdot \frac{D'}{D}\right) \text{Wh/t-km}$$

If no gradient is involved, then specific enrgy consumption is

$$E_{spc} = \left(0.01072 \cdot \frac{V_m^2}{\eta D} \cdot \frac{M_e}{M} + 0.2778 \frac{r}{\eta} \cdot \frac{D'}{D}\right) \text{Wh/t-km}$$

The specific energy consumption of a train running at a given schedule speed is influenced by

1. Distance between stops
2. Acceleration
3. Retardation
4. Maximum speed
5. Type of train and equipment
6. Track configuration.

Adhesive Weight

It is given by the total weight carried on the driving wheels. Its value is $W_a = x\,W$, where W is dead weight and x is a fraction varying from 0.6 to 0.8.

Coefficient of Adhesion

Adhesion between two bodies is due to interlocking of the irregularities of their surfaces in contact. The adhesive weight of a train is ***equal to the total weight to be carried on the driving wheels***. It is less than the dead weight by about 20 to 40%.

If $x = \dfrac{\text{adhesive weight, } W_a}{\text{dead weight } W}$, then, $W_a = x\,W$

Let, F_t = tractive effort to slip the wheels

or

= maximum tractive effort possible without wheel slip

Coefficient of adhesion, $\mu_a = F_t / W_a$

$\therefore$ $F_t = \mu_a W_a = \mu_a x W = \mu_a x Mg$

If M is in tonne, then

$$F_t = 1000 \times 9.8\, x\, \mu_a M = 9800\, \mu_a\, x\, M \text{ newton}$$

It has been found that tractive effort can be increased by increasing the motor torque but only Up to a certain point. Beyond this point, any increase in motor torque does not increase the tractive effort but merely causes the driving wheels to slip. It is seen from the above relation that for increasing F_t, it is not enough to increase the kW rating of the traction motors alone but the weight on the driving wheels has also to be increased.

Adhesion also plays an important role in braking. If braking effort exceeds the adhesive weight of the vehicle, skidding takes place.

Mechanism of Train Movement:

The essentials of driving mechanism in an electric vehicle are illustrated in figure 2.13. The armature of the driving motor has a pinion which meshes with the gear wheel keyed to the axle of the driving wheel. In this way, motor torque is transferred to the wheel through the gear.

Let, T = torque exerted by the motor

F_1 = tractive effort at the pinion

F_t = tractive effort at the wheel

γ = gear ratio

Here, d_1, d_2 = diameters of the pinion and gear wheel respectively

D = diameter of the driving wheel

η = efficiency of power transmission from the motor to driving axle

Now, $T = F_1 \times d_1/2$ or $F_1 = 2T/d_1$

Tractive effort transferred to the driving wheel is

$$F_t = \eta F_1 \left(\frac{d_2}{D}\right) = \eta \cdot \frac{2T}{d_1}\left(\frac{d_2}{D}\right) = \eta T\left(\frac{2}{D}\right)\left(\frac{d_2}{d_1}\right) = 2\,\gamma\,\eta\,\frac{T}{D}$$

For obtaining motion of the train without slipping, $F_t \leq \mu_a W_a$ where μ_a is the coefficient of adhesion and W_a is the adhesive weight.

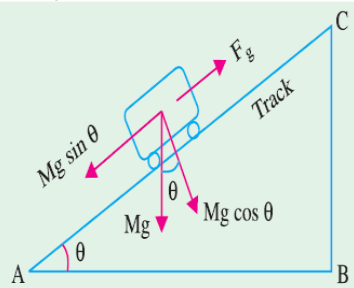


Fig. 2.13

Example:

The peripheral speed of a railway traction motor cannot be allowed to exceed 44 m/s. If gear ratio is 18/75, motor armature diameter 42 cm and wheel diameter 91 cm, calculate the limiting value of the train speed.

Solution. Maximum number of revolutions per second made by armature

$$= \frac{\text{armature velocity}}{\text{armature circumference}} = \frac{44}{0.42\,\pi} = \frac{100}{3}\ \text{rps.}$$

Maximum number of revolutions per second made by the driving wheel

$$= \frac{100}{3} \times \frac{18}{75} = 8\ \text{rps.}$$

Maximum distance travelled by the driving wheel in one second

$$= 8 \times 0.91\,\pi\ \text{m/s} = 22.88\ \text{m/s}$$

Hence, limiting value of train speed

$$= 22.88\ \text{m/s} = 22.88 \times 18/5 = \mathbf{82\ km/h}$$

Example:

A 250-tonne motor coach driven by four motors takes 20 seconds to attain a speed of 42 km/h, starting from rest on an ascending gradient of 1 in 80. The gear ratio is 3.5, gear efficiency 92%, wheel diameter 92 cm train resistance 40 N/t and rotational inertia 10 percent of the dead weight. Find the torque developed by each motor.

Solution. $F_t = (277.8 \times M_e a + 98\ \text{MG} + \text{Mr})$ newton

Now, $\alpha = V_m/t_1 = 42/20 = 2.1$ km/h/s Since gradient is 1 in 80, it becomes 1.25 in 100. Hence, percentage gradient $G = 1.25$. Also, $M_e = 1.1\ M$. The tractive effort at the driving wheel is

$$F_t = 277.8 \times (1.1 \times 250) \times 2.1 + 98 \times 250 \times 1.25 + 250 \times 40$$
$$= 160{,}430 + 30{,}625 + 10{,}000 = 201{,}055\ \text{N}$$

Now, $F_t = 2\gamma\eta T/D$ or $201{,}055 = 2 \times 3.5 \times 0.92 \times T/0.92$ $\therefore T = 28{,}744$ N–m

Torque developed by each motor = 28,744/4 = **7,186 N–m**

Example:

A 250-tonne motor coach having 4 motors, each developing a torque of 8000 N-m during acceleration, starts from rest. If up-gradient is 30 in 1000, gear ratio 3.5, gear transmission efficiency 90%, wheel diameter 90 cm, train resistance 50 N/t, rotational inertia effect 10%, compute the time taken by the coach to attain a speed of 80 km/h.

If supply voltage is 3000 V and motor efficiency 85%, calculate the current taken during the acceleration period.

Solution Tractive effort at the wheel

$$= 2\gamma\,\eta T/D = 2 \times 3.5 \times 0.9 \times (8000 \times 4)/0.9 = 224{,}000\ \text{N}$$

Also, $F_t = (277.8\, a\, M_e + 98\, MG + Mr)$ newton

$$= (277.8 \times (1.1 \times 250)) \times a + 98 \times 250 \times 3 + 250 \times 50\ \text{N}$$
$$= (76{,}395\, a + 86{,}000)\ \text{N}$$

Equating the two expression for tractive effort, we get

$$224{,}000 = 76{,}395\, a + 86{,}000\ ;\ a = 1.8\ \text{km/h/s}$$

Time taken to achieve a speed of 80 km/h is

$$t_1 = V_m / a = 80/1.8 = \mathbf{44.4\ second}$$

Power taken by motors (Art. 41.36) is

$$= \frac{F_t \times v}{\eta} = \frac{F_t \times V_m}{\eta} = F_t \cdot \left(\frac{1000}{3600}\right) \cdot \frac{V_m}{\eta} \text{ watt}$$

$$= 22{,}000 \times 0.2778 \times 80/0.85 = 58.56 \times 10^5 \text{ W}$$

$$\text{Total current drawn} = 55.56 \times 10^5/3000 = 1952 \text{ A}$$

$$\text{Current drawn/motor} = 1952/4 = \mathbf{488\ A.}$$

Example:

A 300-tonne EMU is started with a uniform acceleration and reaches a speed of 40 km/h in 24 seconds on a level track. Assuming trapezoidal speed/time curve, find specific energy consumption if rotational inertia is 8%, retardation is 3 km/h/s, distance between stops is 3 km, motor efficiency is 0.9 and train resistance is 40 N/tonne.

Solution. First of all, let us find *D'* – the distance upto which energy is consumed from the supply. It is the distance travelled upto the end of free-running period. It is equal to the total distance minus the distance travelled during braking.

Braking time, $t_2 = V_m/\beta = 40/3 = 13.33$ second

Distance travelled during braking period

$$= \frac{1}{2} V_m t_3 = \frac{1}{2} \times 40 \times \left(\frac{13.33}{3600}\right) = 0.074 \text{ km}$$

$\therefore \quad D' = D$ – braking distance $= 3 - 0.074 = 2.926$ km

Since, $M_e/M = 1.08$, using the relation derived in Art. 43.43, we get the value of specific energy consumption as

$$= \left(0.01072 \frac{V_m^2}{\eta D} \cdot \frac{M_e}{M} + 0.2778 \frac{r}{\eta} \frac{D'}{D}\right) \text{Wh/t-km}$$

$$= \left(0.01072 \times \frac{40^2}{0.9 \times 3} \times 1.08 + 0.2778 \times \frac{49}{0.9} \times \frac{2.926}{3}\right) = \mathbf{21.6\ Wh/t\text{-}km.}$$

Example:

An electric train accelerates uniformly from rest to a speed of 50 km/h in 25 seconds. It then coasts for 70 seconds against a constant resistance of 60 N/t and is then braked to rest with uniform retardation of 3.0 km/h/s in 12 seconds. Compute (i) uniform acceleration (ii) coasting retardation (iii) schedule speed if station stops are of 20-second duration

Allow 10% for rotational inertia. How will the schedule speed be affected if duration of stops is reduced to 15 seconds, other factors remaining the same?

Solution. *(i)* As seen from Fig. 43.15, $\alpha = V_1/t_1 = 50/25 = $ **2 km/h/s**

(ii) The speeds at points *B* and *C* are connected by the relation

$$0 = V_2 + \beta t_3 \quad \text{or} \quad 0 = V_2 + (-3) \times 12 \qquad \therefore \quad V_2 = 36 \text{ km/h}$$

Coasting retardation, $\beta_c = (V_2 - V_1)/t_2 = (36 - 50)/70 = -$ **0.2 km/h/s** *(iii)* Distance travelled during acceleration

$$= \frac{1}{2} V_1 t_1 = \frac{1}{2} \times 50 \frac{\text{km}}{\text{h}} \times \frac{25}{3600} \text{h}$$

$$= 0.174 \text{ km}$$

Distance travelled during coasting can be found from the relation

$$V_{22} - V_{12} = 2\ \beta_c D \quad \text{or}$$

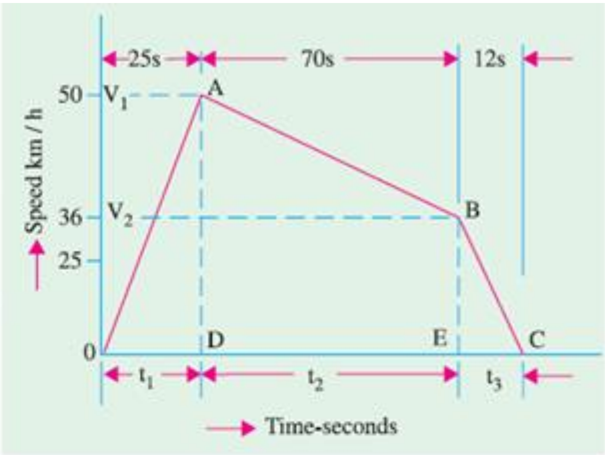
$$D = (362 - 502)/2 \times -0.2 \times 3600$$

$$= 0.836 \text{ km}$$

Distance covered during braking

$$= \frac{1}{2} V_2 t_3 = \frac{1}{2} \times 36 \frac{\text{km}}{\text{h}} \times \frac{12}{3600} \text{h}$$

$$= 0.06 \text{ km}$$



Total distance travelled from start to stop

$$= 0.174 + 0.836 + 0.06 = 1.07 \text{ km}$$

Total time taken including stop time

$$= 25 + 70 + 12 + 20 = 127 \text{ second}$$

Schedule speed $= 1.07 \times 3600/127 =$ **30.3 km/s**

Schedule speed with a stop of 15 s is $= 1.07 \times 3600/122 =$ **31.6 km/h**

General Features of Traction Motor:

Electric Features

- High starting torque
- Series Speed - Torque characteristic
- Simple speed control
- Possibility of dynamic/ regenerative braking
- Good commutation under rapid fluctuations of supply voltage.

Mechanical Features

- Robustness and ability to withstand continuous vibrations.
- Minimum weight and overall dimensions
- Protection against dirt and dust

No type of motor completely fulfils all these requirements. Motors, which have been found satisfactory, are D.C. series for D.C. systems and A.C. series for A.C. systems. While using A.C. three phase motors are used. With the advent of Power Electronics it is very easy to convert single phase A.C. supply drawn from pantograph to three phase A.C.

Speed - Torque Characteristic of D.C. Motor:

$$V = E_b + I_a R_a$$

$$V \,.\, I_a = E_b \,.\, I_a + I_a^2 R_a$$

where E_b I_a = Power input to armature = Electrical power converted into mechanical power at the shaft of motor.

$$Mechanical\ Power = T.\omega = T \times \frac{2\pi N}{60}$$

$$\therefore \frac{2\pi NT}{60} = E_b.I_a$$

$$\therefore T = \frac{60E_b.I_a}{2\pi N} = 9.55\frac{E_b.I_a}{N}$$

$$But\ E_b = \frac{\phi ZNP}{60A}$$

$$\therefore \quad T = 9.55\frac{\phi ZNP}{60\ A}\frac{I_a}{N} = 9.55\frac{\phi ZP}{60}\frac{I_a}{A}$$

$$= 0.1592 \times \phi \times \left[Z\frac{I_a}{A}\right] P \text{ Nw-m}$$

∴ Torque $T = 0.1592 \times$ flux per pole $\times$ armature amp. conductors $\times$ Number of poles

Also speed 'N' can be calculated as:

$$E_b = \frac{\phi ZNP}{60\ A} \qquad \therefore \quad N = \frac{(E_b)}{\phi ZP} 60\ A$$

$$N = \frac{(V - I_a R_a)\ 60\ A}{\phi ZP} \qquad \therefore \quad N \propto \frac{V - I_a R_a}{\phi}$$

But $T = 9.55\frac{\phi ZP}{60}\frac{I_a}{A}$ from the equation of torque

$$\therefore \quad \frac{T}{I_a} = \frac{9.55\ \phi ZP}{60\ A} \Rightarrow \frac{9.55\ I_a}{T} = \frac{60\ A}{\phi ZP}$$ Put this value in the above equation of N

$$\therefore \quad N = \frac{(V - I_a R_a) \times 9.55\ I_a}{T}$$

Speed $$\boxed{N = \frac{9.55\ (V - I_a R_a)}{T / I_a}}$$

The torque - current and speed - torque curves for D.C. motors are shown in figure 2.14 (a) and (b) respectively.

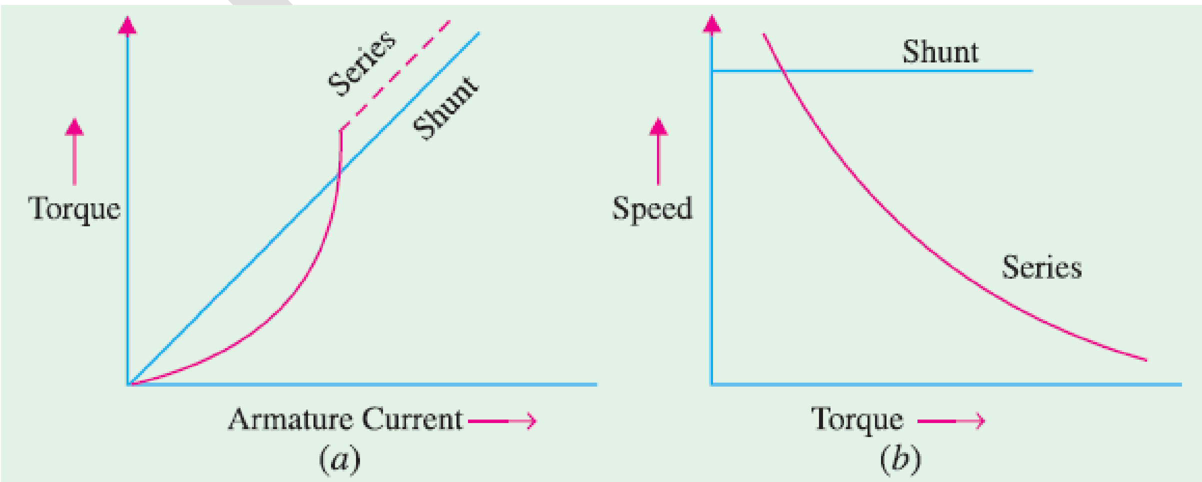


Fig. 2.14

Control of D.C. Motors:

The starting current of motor is limited to its normal rated current by starter during starting. At the instant of switching on the motor, back e.m.f. $E_b = 0$

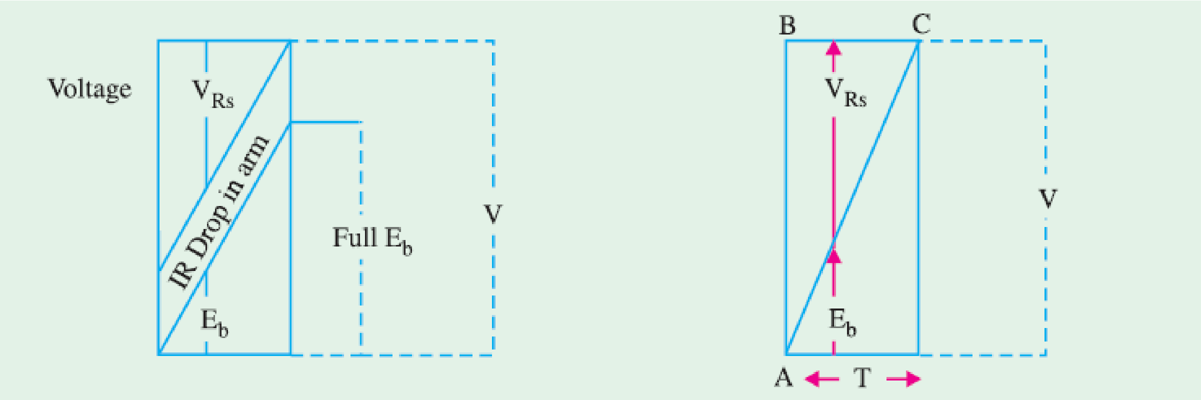
∴ Supply voltage = $V = IR$ + Voltage drop across R_s.

At any other instant during starting

$V = IR$ + Voltage across $R_s + E_b$

At the end of accelerating period, when total R_s is cut-off

$V = E_b + IR$



(*a*) Voltage during starting of D.C. motor (*b*) Voltage during starting neglecting IR drop

Fig. 2.15

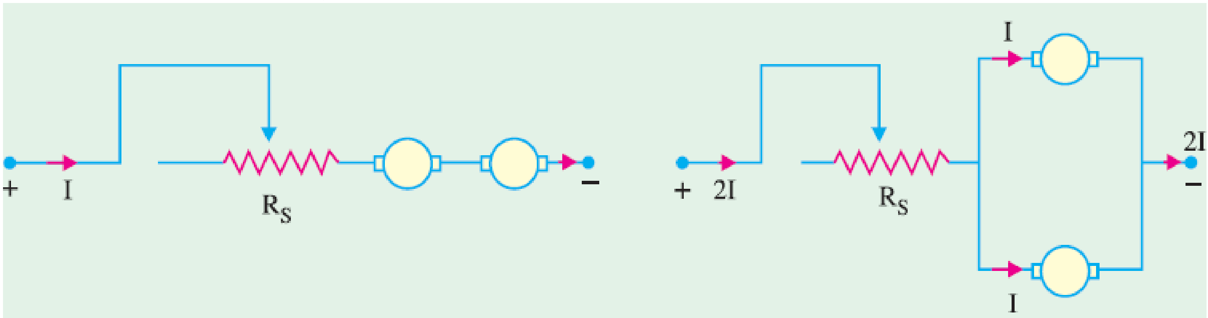
If T is the time in sec. for starting and neglecting IR drop, total energy supplied = $V.I.T.$ watt-sec From figure 2.15 (*b*) Energy wasted in R_s = Area of triangle $ABC \times I$ = ½. $T.V.I.$ watt - sec. = ½ VIT watt - sec. But total energy supplied = $V.I.T$ watt - sec.

∴ Half the energy is wasted in starting

∴ $\eta_{starting} = 50\%$

Series - Parallel Starting:

With a 2 motor equipment ½ the normal voltage will be applied to each motor at starting as shown in figure 2.16 (*a*) (Series connection) and they will run upto approximate ½ speed, at which instant they are switched on to parallel and full voltage is applied to each motor. R_s is gradually cutout, with motors in series connection and then reinserted when the motors are connected in parallel, and again gradually cut-out.



(*a*) Motors in series (*b*) Motors in parallel

Fig. 2.16

In traction work, 2 or more similar motors are employed. Consider 2 series motors started by series parallel method, which results in saving of energy.

(a)Series operation. The 2 motors, are started in series with the help of R_s. The current during starting is limited to normal rated current 'I' per motor. During series operation, current 'I' is drawn from supply. At the instant of starting OA = AB = IR drop in each motor. OK = Supply voltage 'V'.

The back e.m.fs. of 2 motors jointly develop along OM as shown in figure 4.16 (a). At point. E, supply voltage V = Back e.m.fs of 2 motors + IR drops of 2 motor. Any point on the line BC represents the sum of Back e.m.fs. of 2 motors + IR drops of 2 motors + Voltage across resistance R_s of 2 motors OE = time taken for series running.

At pt 'E' at the end of series running period, each motor has developed a back e.m.f.

$$= \frac{V}{2} - IR$$

$$EL = ED - LD$$

(b) **Parallel operation.** The motors are switched on in parallel at the instant 'E', with R_s reinserted as shown in figure 2.16(b). Current drawn is $2I$ from supply. Back e.m.f. across each motor = EL. So the back e.m.f. now develops along LG. At point 'H' when the motors are in full parallel, (R_s = 0 and both the motors are running at rated speed)

Supply voltage = $V = HF = HG + GF$

= Normal Back e.m.f. of each motor + IR drop in each motor.

To find t_s, t_p and η of starting:

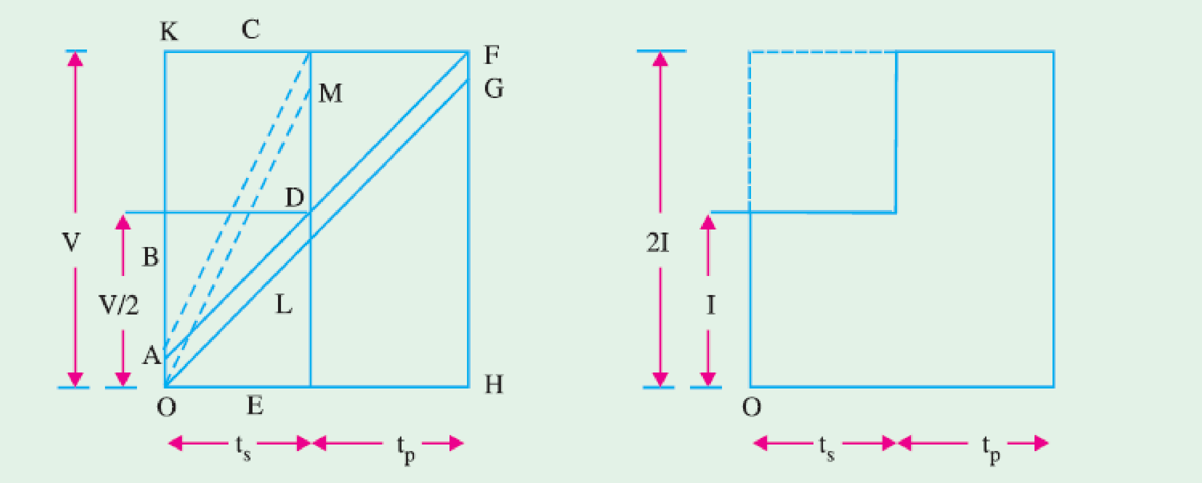
The values of time t_s during which the motors remain in series and t_p during which they are in parallel can be determined from figure 2.17(a), (c). From figure 2.17(a), triangles OLE and OGH are similar

$$\therefore \quad \frac{OE}{OH} = \frac{LE}{GH} \quad \therefore \frac{t_s}{T} = \frac{DE - DL}{FH - FG} = \frac{V/2 - IR}{V - IR}$$

$$\therefore \quad t_s = \frac{1}{2}\left(\frac{V - 2IR}{V - IR}\right)T$$

$$t_p = T - t_s = T - \left\{\frac{1}{2}\left(\frac{V - 2IR}{V - IR}\right)T\right\}$$

$$t_p = T\left\{1 - \frac{1}{2}\left(\frac{V - 2IR}{V - IR}\right)T\right\}$$



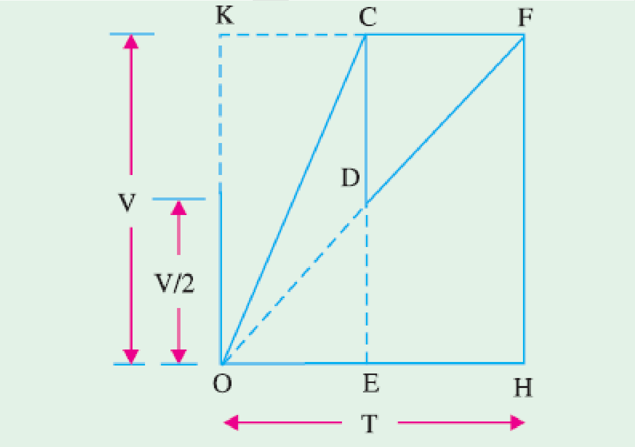
(*a*) Voltage built-up in series-parallel starting (*b*) Variation of current in series-parallel starting

Fig. 2.17

To calculate η of starting, neglect *IR* drop in armature circuit.

This modifies figure 2.17 (*a*) to figure 2.17 (*c*). '*D*' is midpoint of *CE* and back e.m.f. develops along *DF* in parallel combination. *KC* = *CF i.e.* time for series combination = time for parallel combination

i.e. $t_s = t_p = t$ and average starting current = *I* per motor.



(*c*) Efficiency of starting by series-parallel method

Fig. 2.17

Energy lost in R_s = Area under triangle *OKC* + Area under triangle *CDF*

$$= \left(\frac{1}{2} V\, I\right) \times t + \left(\frac{1}{2}\, \frac{V}{2}\, 2I\right) \times t = VIt$$

But total energy supplied

$$= \underset{\text{(Series)}}{I\,V\,t} + \underset{\text{(Parallel)}}{2\,I\,V\,t}$$

$$= 3\,V\,I\,t$$

$$\therefore \eta \text{ of starting} = \frac{3VIt - VIt}{3VIt}$$

$$= \frac{2}{3} = 66.6\%$$

∴ η is increased by 16.66% as compared to previous case. If there are 4 motors then $\eta_{starting}$ = 73%. So there is saving of energy lost in R_s, during starting period as compared with starting by both motors in parallel.

Series Parallel Control by Shunt Transition Method:

The various stages involved in this method of series – parallel control are shown in figure 2.18. In steps 1, 2, 3 and 4 the motors are in series and are accelerated by cutting out the R_s in steps. In step 4, motors are in full series. During transition from series to parallel, R_s is reinserted in circuit– step 5. One of the motors is bypassed -step 6 and disconnected from main circuit – step 7. It is then connected in parallel with other motor -step 8, giving 1st parallel position. R_s is again cut-out in steps completely and the motors are placed in full parallel.

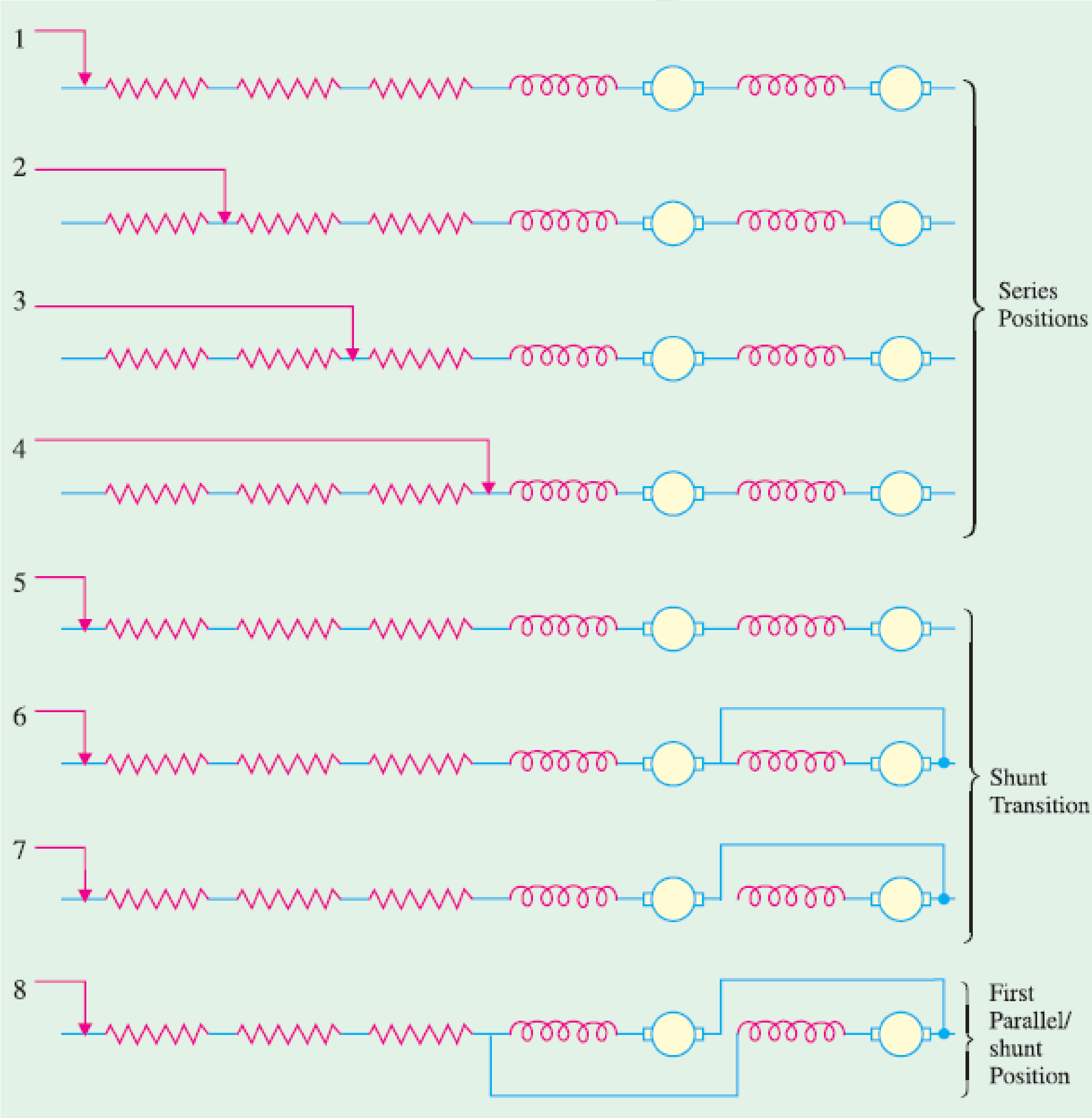


Fig. 2.18

The main difficulty with series parallel control is to obtain a satisfactory method of transition from series to parallel without interrupting the torque or allowing any heavy rushes of current. In shunt transition method, one motor is short circuited and the total torque is reduced by about 50% during transition period, causing a noticeable jerk in the motion of vehicle.

The Bridge transition is more complicated, but the resistances which are connected in parallel with or 'bridged' across the motors are of such a value that current through the motors is not altered in magnitude and the total torque is therefore held constant and hence it is normally used for railways. So in this method it is seen that, both motors remain in circuit through-out the transition. Thus the jerks will not be experienced if this method is employed.

Series Parallel Control by Bridge Transition:

(*a*) At starting, motors are in series with R_s *i.e.* link *P* in position = $AA^{'}$

(*b*) Motors in full series with link *P* in position = $BB^{'}$ (No R_s in the circuit)

The motor and R_s are connected in the form of Wheatstone Bridge. Initially motors are in series with full R_s as shown in figure 2.19(*a*). *A* and $A^{'}$ are moved in direction of arrow heads. In position $BB^{'}$ motors are in full series, as shown in figure 2.19(*b*), with no R_s present in the circuit. In transition step the R_s is reinserted.

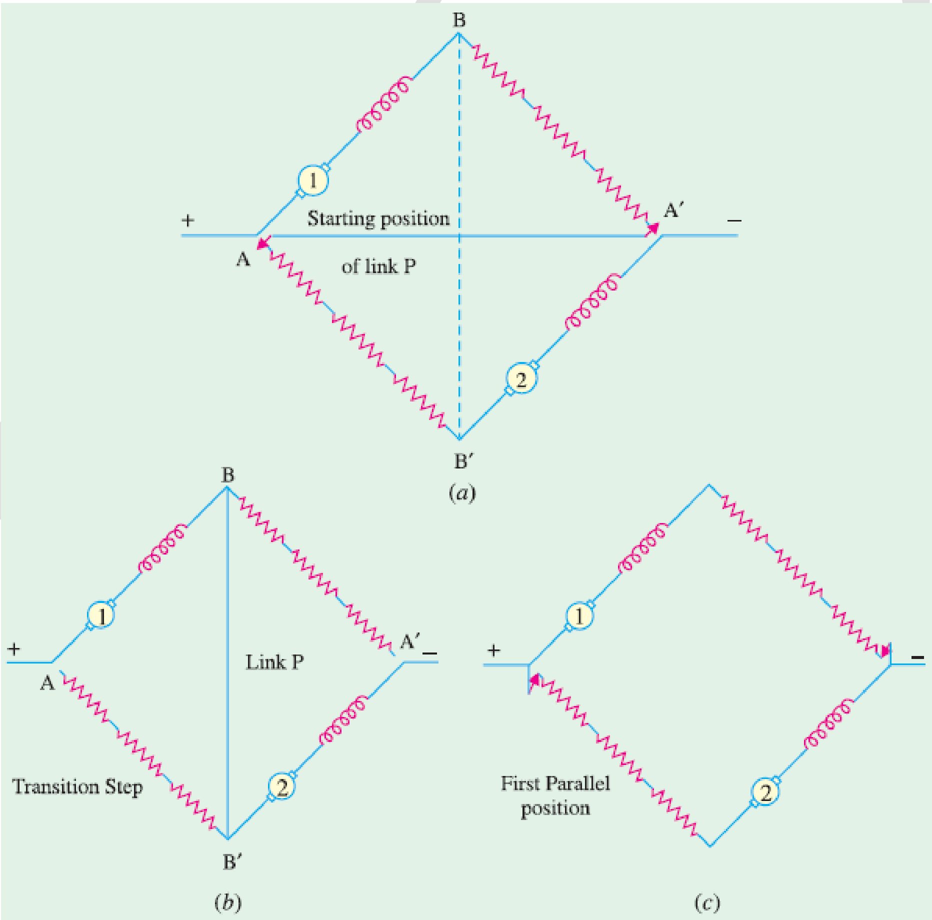


Fig. 2.19

In Ist parallel step, link *P* is removed and motors are connected in parallel with full R_s as shown in figure 2.19 (*c*). Advantage of this method is that the normal acceleration torque is available from both the motors, through - out starting period. Therefore acceleration is smoother, without any jerks, which is very much desirable for traction motors.

Questions Bank

1. What do you mean by Electrolysis process?
2. Explain the extraction of metals.
3. Define electro chemical equivalent and energy efficiency.
4. State and explain faradays laws of electrolysis.
5. Explain throwing power and polarisation.
6. What do you mean by electro deposition?
7. Explain the factors affecting electro deposition.
8. What are the requirements of an ideal traction system?
9. What are the advantages and disadvantages of electric traction?
10. Draw the speed-time curve of a main line service and explain.
11. What are the various electric traction systems in India? Compare them.
12. Give the features of various motors used in electric traction.
13. Draw the speed-time curve of a suburban service train and explain.
14. Derive an expression for specific energy output on level track using a simplified speed-time curve. What purpose is achieved by this quantity?
15. Explain the characteristics of series motors and also explain how they are suitable of electric traction work?
16. For a trapezoidal speed-time curve of a electric train, derive expression for maximum speed and distance between stops.
17. Derive expression for the tractive effort for a train on a level track.
18. What are various types of traction motors?
19. For a quadrilateral speed-time curve of an electric train, derive expression for the distance between stops and speed at the end of the coasting period.
20. Discuss why a D.C. series motor is ideally suited for traction services.
21. State factors affecting specific energy consumption.
22. Explain the general features of traction motors.
23. Explain the terms (i) tractive effort (ii) coefficient of adhesion (iii) specific energy consumption of train (iv) tractive resistance.
24. Briefly explain the controlling of D.C. Motor.
25. Find the schedule speed of an electric train for a run of 1.5 km if the ratio of its maximum to average speed is 1.25. It has a braking retardation of 3.6 km/h/s, acceleration of 1.8 km/h/s and stop time of 21 second. Assume trapezoidal speed/time curve.
26. A train runs between two stations 1.6 km apart at an average speed of 36 km/h. If the maximum speed is to be limited to 72 km/h, acceleration to 2.7 km/h/s, coasting

retardation to 0.18 km/h/s and braking retardation to 3.2 km/h/s, compute the duration of acceleration, coasting and braking periods. Assume a simplified speed/time curve.

27. The peripheral speed of a railway traction motor cannot be allowed to exceed 44 m/s. If gear ratio is 18/75, motor armature diameter 42 cm and wheel diameter 91 cm, calculate the limiting value of the train speed.

28. A 250-tonne motor coach driven by four motors takes 20 seconds to attain a speed of 42 km/h, starting from rest on an ascending gradient of 1 in 80. The gear ratio is 3.5, gear efficiency 92%, wheel diameter 92 cm train resistance 40 N/t and rotational inertia 10 percent of the dead weight. Find the torque developed by each motor.

29. A 250-tonne motor coach having 4 motors, each developing a torque of 8000 N-m during acceleration, starts from rest. If up-gradient is 30 in 1000, gear ratio 3.5, gear transmission efficiency 90%, wheel diameter 90 cm, train resistance 50 N/t, rotational inertia effect 10%, compute the time taken by the coach to attain a speed of 80 km/h. If supply voltage is 3000 V and motor efficiency 85%, calculate the current taken during the acceleration period.

30. A goods train weighing 500 tonne is to be hauled by a locomotive up an ascending gradient of 2% with an acceleration of 1 km/h/s. If coefficient of adhesion is 0.25, train resistance 40 N/t and effect of rotational inertia 10%, find the weight of locomotive and number of axles if load is not to increase beyond 21 tonne/axle.

31. The average distance between stops on a level section of a railway is 1.25 km. Motor-coach train weighing 200 tonne has a schedule speed of 30 km/h, the duration of stops being 30 seconds. The acceleration is 1.9 km/h/s and the braking retardation is 3.2 km/h/s. Train resistance to traction is 45 N/t. Allowance for rotational inertia is 10%. Calculate the specific energy output in Wh/t-km. Assume a trapezoidal speed/time curve.

32. An electric train has a quadrilateral speed-time curve as follows: (i) uniform acceleration from rest at 2.5 km/h/s for 25 second (ii) coasting for 50 second (iii) duration of braking 25 second. If the train is moving along a uniform upgradient of 1 in 100 with a tractive resistance of 45 N/t, rotational inertia 10% of dead weight, duration of stops at stations 20 second and overall efficiency of transmission gear and motor 80%, calculate the schedule speed and specific energy consumption of run. **[69 km/h, 26.61 Wh/t-km]**

33. A train weighing 400 tonne has speed reduced by regenerative braking from 40 to 20 km/h over a distance of 2 km at a down gradient of 20%. Calculate the electrical energy and average power returned to the line. Tractive resistance is 40 N/t and allow rotational inertia of 10% and efficiency of conversion 75%. **[324 kW/h, 4860 kW]**

UNIT-3

ELECTRIC TRACTION

AC Traction Equipment:

The figure below shows the block diagram of an AC locomotive system that employs single phase supply to drive three phase motor. The various components of this system include overhead contact wire, circuit breakers, pantograph, transformer, three phase traction motor, rectifier, inverter, smoothing reactor, etc

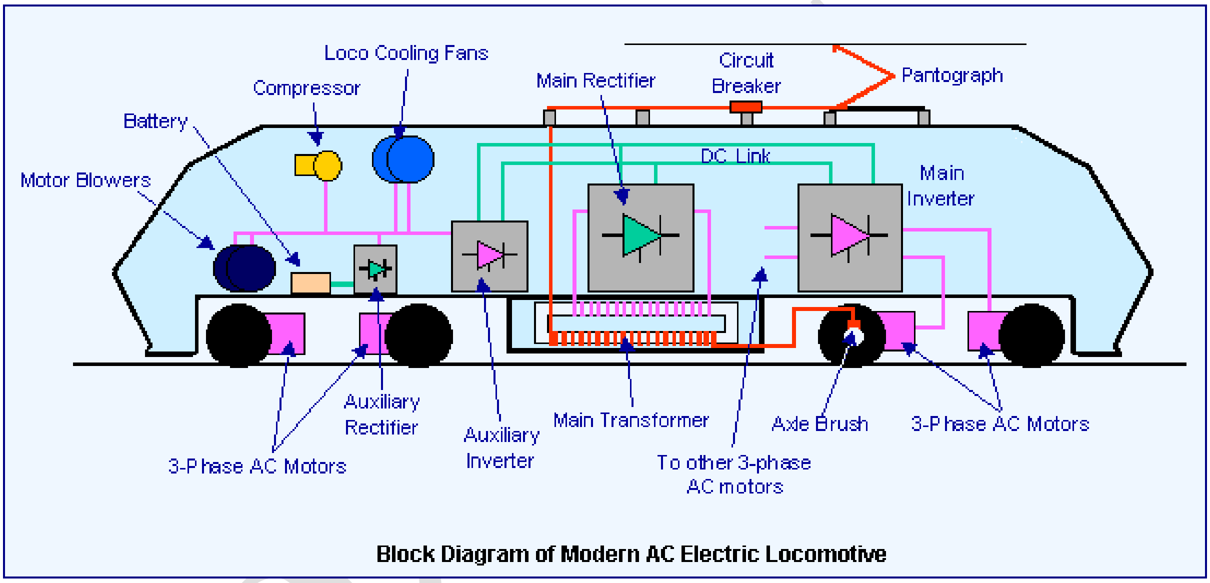


Fig. 3.1

Pantograph: The main function of pantograph is to maintain link between overhead conductor and power circuit of locomotive at different speeds of the vehicle under all wind conditions. It collects the current from overhead conductor and supplies to rest circuit.

Circuit Breaker: It protects the power circuit in the event of any fault by isolating it from the supply. It also isolates the circuit during maintenance.

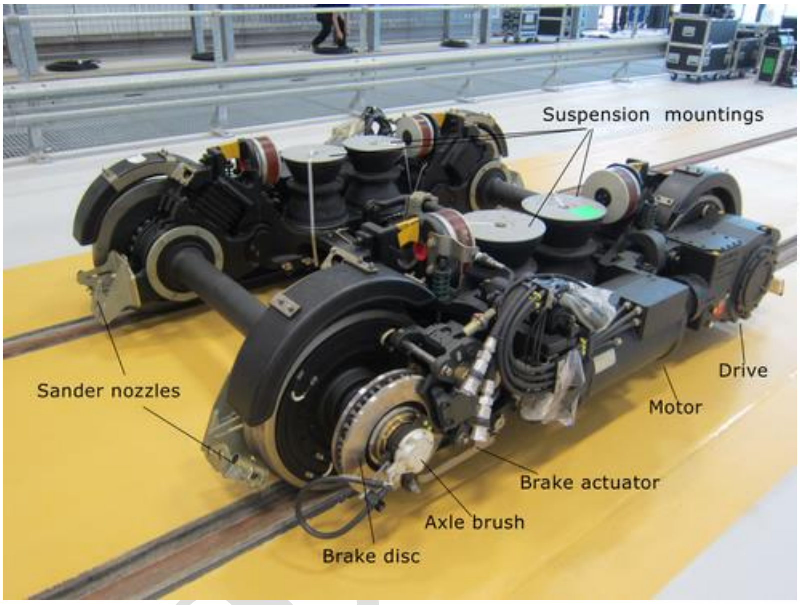
Transformer: It receives the high voltage from overhead conductor via pantograph and circuit breaker and then step-down the voltage to desired level required by the rest circuit.

Rectifier: It converts a low voltage AC supply from the secondary of transformer to a DC supply.

DC Link: It connects the rectifier and inverter circuits. It consists of filter arrangement (capacitor and inductor arrangement) that filters the output from rectifier (by removing the harmonics form it) and then supplies it to the inverter.

Main Inverter: It converts the DC power to three phase AC power in order to drive three phase AC motors.

Axle Brush: It acts as a return path for the supply. Once the power is drawn to the locomotive from overhead system, the current complete its path through axle brush and one of running tacks.



Auxiliary Inverter: This inverter supplies the power to other parts in the locomotive unit including fans, motor blowers, compressors, etc.

Battery: It supplies the necessary starting current and also power up the essential circuits such as emergency lighting.

Compressor: It maintains the cooling/heating requirement in the locomotive unit.

Cooling Fans: These fans maintain the necessary cooling for the power circuits. Modern locomotive systems use electronically controlled air management systems to keep the desired temperature.

AC Series Motors:

Many single phase ac motors have been developed for traction purposes but only compensated series type commutator motor is found to be best suited for traction. Single phase induction motors have been abandoned as they are not capable of developing high starting torque.

The construction of an AC series motor is very similar to a DC series motor except that some modifications (such as whole magnetic circuit laminated, series field with as few turns as possible, large number of armature conductors, use of high resistance carbon brushes, numerous poles with lesser flux per pole, very short air gaps etc.) are incorporated s as to obtain better performance. Compensating windings are provided to neutralize armature reaction and commutating or interpoles are provided for better performance in terms of higher efficiency and a greater output from a given size of armature core.

A schematic diagram of a single phase series motor with interpole and compensating windings is illustrated in figure 3.2.

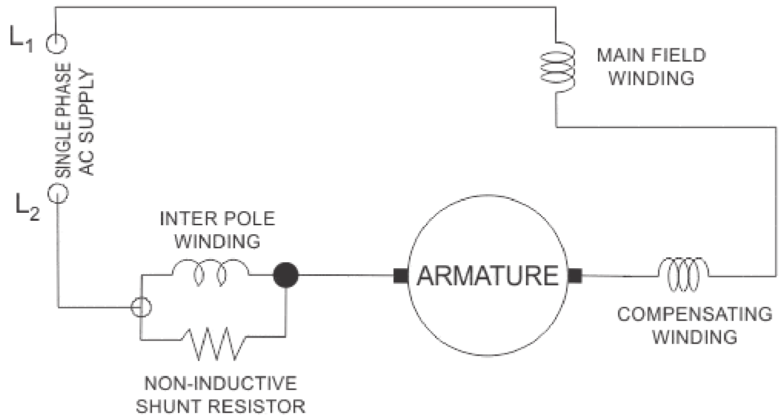


Fig. 3.2. AC series motor with interpoles and compensating windings

The average value of the torque on the motor shaft is given as

$$T_{av} = \frac{1}{2\pi\sqrt{2}} Z \frac{P}{A} \phi_{max} I \cos\theta$$

where, I is the effective value of current, ϕ_{max} is the peak value of the flux per pole and θ is the phase angle between phasors ϕ and I.

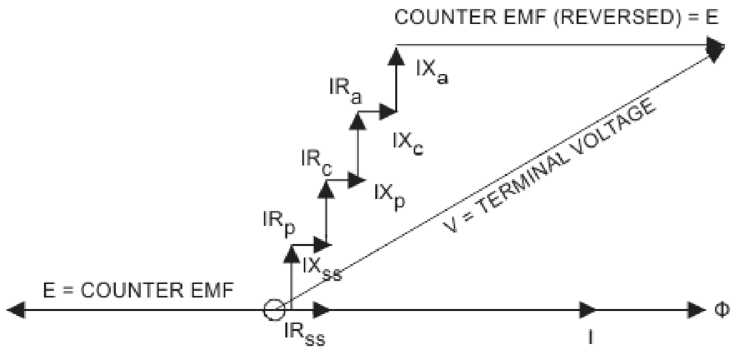


Fig. 3.3 Phasor Diagram for a Single Phase Series Motor

The phasor diagram of a single phase ac series motor is given in figure 3.3. Flux ϕ is produced by current I. Flux ϕ lags behind the current I by an amount too small to be important, so it has been shown in phase with current I. The resistive drops IR_{ss}, IR_p, IR_c and IR_a due to resistances of series field, interpole circuit, compensating winding and of armature respectively are in phase with current I. The reactive drops IX_{ss}, IX_p, IX_c and IX_a due to reactances of series field, interpole circuit, compensating winding and of armature respectively are in quadrature with current I and leading. Here it is noteworthy that reactive drop due to series field is much greater than that of either armature or the compensating field even all the measures to reduce series field inductance, are taken. Unlike the armature cross-flux ϕ', the flux that causes this reactive drop cannot be neutralized because it is the necessity for the development of torque. The interpole circuit consists of the interpole winding in parallel with the non-inductive shunt, as illustrated in figure 3.2. R_p and X_p are the respective equivalent resistance and reactance of the interpole parallel circuit.

The phase angle of E, in the phasor diagram is most easily determined by considering instantaneous values. When the alternating field flux ϕ is at its peak value, the armature conductors cut the maximum flux, and the rotational or speed emf is therefore a maximum. When the field flux is zero, the rotational or speed emf, E is zero. Since the graphs of flux ϕ and generated emf E pass through zero at the same instant and also through maximum at the same instant, the phasor E must be either exactly in phase with the phasor of flux ϕ or exactly in phase opposition. The first possibility is readily eliminated by noting that in case of a motor the generated voltage is in opposition to the direction of flow of current, and, therefore, E is in phase opposition with current and is called the counter emf.

The terminal voltage V is the phasor sum of the counter- emf reversed and all the resistive and reactive volt drops.

Series Motor Characteristics:

1. ***Power-Output Characteristic:*** Mechanical power developed is given by the product of counter emf E and the armature current I. Though counter emf E decreases slightly with the increase in current, but if it is neglected in comparison with the magnitude of counter emf E, mechanical power developed increases almost in proportion with current I.

Power available at shaft or power output is equal to the mechanical power developed less rotational losses. Power output characteristic is shown in figure 3.4.

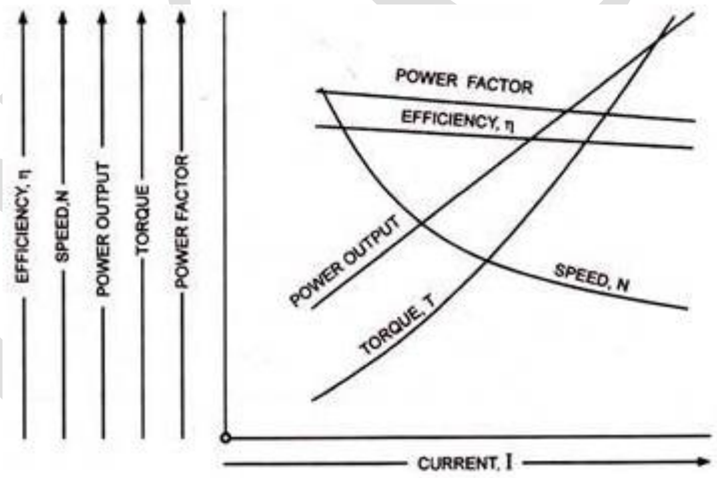


Fig. 3.4

2. ***Power Factor Characteristic:*** The cosine of the phase angle ϕ is the power factor of the motor.

From phasor diagram shown in figure 3.3

$$sin\phi = I \times \frac{X_s + X_p + X_c + X_a}{V} = I\frac{X}{V}$$

where $X = X_s + X_p + X_c + X_a$ = effective reactance of the motor

or Power factor, $cos\phi = \sqrt{1 - \left(\frac{IX}{V}\right)^2}$

In order to have high power factor the reactance drops must be low and counter emf high. The reactance drops are lowest and the counter emf is highest at light loads. And therefore, the power factor of the single phase series motor is highest (nearly unity) at light loads and falls with the increase in current. This is the reason that power factor of series motor is low at overloads and at starting. This is the reverse of the power factor relation that exists in case of the induction motor and transformer.

3. ***Speed-Current Characteristic:*** Speed of the commutator motor is proportional to counter emf E (E_r or E_b) or proportional to motor supply voltage less voltage drops because

$$E \propto \phi N \text{ and } E = V - \text{voltage drops}$$

The current causes resistive drops in series field and armature in case of dc operation but in case of ac operation it also causes large reactance drops in series field, armature, compensating winding and interpole circuit in addition to resistive drops in these windings. Hence with ac operation, counter emf E developed is much less than that with dc operation. It means speed-current characteristic for ac series motors are more drooping as compared to those for dc series motors, as illustrated in figure 3.5.

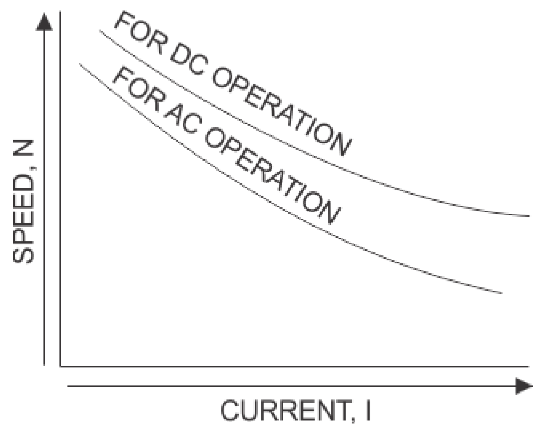


Fig. 3.5

Let the same series commutator motor be operated on dc as well as on ac and with same current I.

The back emf developed in case of dc operation is given as

$$E_b = \frac{\phi Z N_{dc}}{60} \times \frac{P}{A} \text{ volts}$$

and

$$E_r = \frac{\phi_{max}}{\sqrt{2}} \frac{Z N_{ac}}{60} \times \frac{P}{A} \text{ volts}$$

Neglecting saturation, $\phi_{max} = \sqrt{2}\,\phi$ as the peak value of ac flux ϕ_{max} is set up by the peak value of current $\sqrt{2}$ while dc flux ϕ is set up by current I. So from above two equations

$$\frac{N_{ac}}{N_{dc}} = \frac{E_r}{E_b} = \frac{V\cos\phi - IR}{V - IR} = \cos\phi$$

If the supply voltage is same in both of the cases and resistive drops IR are neglected in comparison with supply voltage V, where ϕ is the phase angle between supply voltage and current.

4. Torque-Current Characteristic: From torque equation $T\alpha\ \phi I$, if phase angle between flux ϕ and current I is neglected or $T\alpha\ I^2$ neglecting saturation. Under saturated condition, flux ϕ becomes constant and, therefore, torque becomes directly proportional to current l. Torque-current characteristic for an AC series motor is shown in figure 3.6.

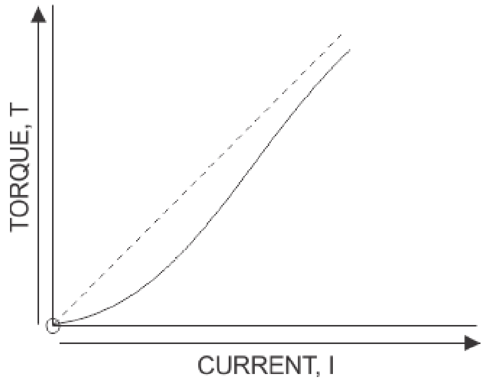


Fig. 3.6

***5. Torque-Speed Characteristic*:** The torque-speed characteristic for an ac series motor can be derived from speed- current and torque-current characteristics given in figures 3.5 and 3.6 respectively.

For given value of torque T and applied voltage V, the armature current is same but voltage drop in case of ac series motor is more than that in case of a dc series motor . Speed of an ac series motor for a given developed torque is less than that of a dc series motor, as illustrated in figure 3.7.

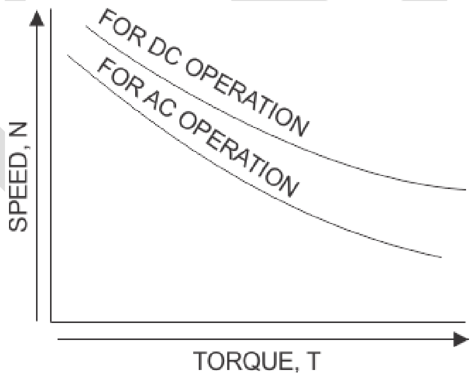


Fig. 3.7

The single phase ac series motor has practically the same operating characteristics as the dc series motor. The torque, or tractive effort, varies nearly as the square of the current and the speed varies inversely as the current. This is illustrated in figure 3.4.

However, in case of an ac series motor (*i*) power factor is very low at starting and on overloads on account of high, inductive nature of the series field and armature circuits (*ii*) efficiency is not as good as in a corresponding dc machine due to eddy current losses and effects of power factor and (*iii*) starting torque is low due to poor power factor at starting.

For a given kilowatt rating ac series motor is 1.5-2 times in size and weight of the corresponding dc series motor. The construction cost of an ac series motor is much more than a dc series motor.

The speed of an ac series motor may be controlled efficiently by taps on a transformer, which is not possible in case of a dc series motor.

The torque-speed characteristic of the single phase series motor is similar to that of the dc series motor. i.e., high starting torque and decrease in speed with increase in load making it to have self-relieving property from heavy excessive load, so such a machine is particularly useful for traction services.

These motors are used for main line services. These are not suitable for urban and suburban services because of low starting torque and poor power factor at start.

However, single phase series motors have better performance (improved power factor, higher efficiency, improved commutation, better starting and fewer poles for a given output) on reduced supply frequency (say $16\frac{2}{3}$ or 25 Hz). The higher frequency results in higher leakage reactances and hence a relatively poor power factor. The weight per kW is also greater for the higher frequency because of dimensions of the motor.

The 50 Hz motors have steeper speed-torque characteristics as they operate at lower flux densities than a $16\frac{2}{3}$ or 25 Hz motors. This results in fewer voltage steps at starting and the division of load between parallel connected motors will be less affected by wear of the driving wheels.

The operating voltage is kept low (300 to 400 V) in order to reduce the inductance. Because high voltage motor with proportionately low current would require a large number of turns to produce the given flux.

Three Phase Induction Motors:

The three phase induction motors have the advantages of simple and robust construction, trouble free operation, less I maintenance, high voltage operation consequently requiring reduced amount of current and automatic regeneration. But due to their flat speed-torque characteristics, constant speed operation, developing low starting torque, drawing high starting H current, complicated speed control system and complicated overhead feeding systems they are not suitable for electric traction work. It is also impossible to employ 3-phase induction motors for a multiple unit system in which two or more locomotives are used for propelling a heavy train. In traction system employing 3- phase induction motors it becomes necessary to couple all the driving axles through a connecting rod so that there is no possibility of a difference in speeds.

With the development of thyristorised inverter circuits, it has now been possible to invert the supply and obtain a variable frequency supply which could be used for the 3- phase induction motor and a very smooth speed control can be obtained.

Induction motor gives good power factor and good efficiency at the speed near to the synchronous speed of the motor. Now in conventional methods of speed control, change of slip method is normally used. But in this method at low speeds of motor or at high value of slip, rotor losses will be more so good efficiency cannot be obtained. To get good, efficiency and good power factor, synchronous speed of the motor itself can be brought to lower speed near to the desired actual speed of the motor. This can be done by supplying the motor with the help of variable frequency supply.

Starting current of motor also decreases when motor is started at low frequency. When 3-phase induction motor is used, distribution system consists of two overhead wires and track rail for the third phase to feed power to locomotive. This gives a complicated overhead structure and also a person who comes in contact of third live rail might be in danger. This drawback is removed by employing Kando system. In this system single phase HV supply is given to locomotive with the help of single overhead wire. In locomotive, this single phase supply is converted into three phase supply with the help of phase converters and fed to the motor. Kando system is in use in Hungary and in some sections of Italian State Railway.

Linear Induction Motors:

It is a special type of induction motor which gives linear motion instead of rotational motion as in the case of a conventional induction motor. It operates on the same principle on which a conventional induction motor operates *i.e.*, "whenever there occurs a relative motion between the field and the short-circuited conductors, currents are induced in them which results in electromagnetic forces and under the influence of these forces, according to Lenz's law, the conductors try to move in such a way as to eliminate the induced currents". In case of a conventional induction motor, movement of field is rotary about an axis so the movement of the conductors is also rotary. But is case of a linear induction motor, the movement of the field is rectilinear and so the movement of conductors.

In its simplest form, a linear induction motor consists of field system having a 3-phase distributed winding placed in slots as shown in figure 3.8. The field system may be a single primary system or double primary system (Fig. 3.8). The secondary of the linear induction motor is normally a conducting plate made of either copper or aluminum in which interaction currents are induced. Either member can be the *stator*, the other being the *runner* in accordance with the particular requirements imposed by the duty for which motor is intended.

In a single primary system a ferromagnetic plate is usually placed on the other side of the conducting plate to provide a path of low reluctance to the main flux. However, the ferromagnetic plate gets attracted towards the primary on energisation of the field and this causes unequal gap length on two sides of the conducting plate. This problem can be overcome by employing double primary system. [Fig. 3.8 (b)].

Which of the two primary and secondary will be shorter in length compared to the other depends upon the use of the motor. When the operating distance is large, the primary is made shorter than the secondary because it is uneconomical to wind a very long 3-phase primary. The short secondary form [Fig. 3.8 (c)] is useful with limited operating distance.

When the three phase primary winding of a linear induction motor is energized from a balanced three phase source, a magnetic field moving in a straight line from one end to the other at a linear synchronous speed v_s is developed. The linear synchronous speed is given as

$$v_s = 2\tau f \text{ meters/second}$$

where τ is the pole pitch in meters and f is the supply frequency in Hz.

It is to be noted here that the synchronous speed does not depend on the number of poles, but only on the pole pitch and the stator supply frequency.

As the flux moves linearly, it drags rotor plate along with it in the same direction. This reduces the relative speed of travel of the flux with respect to rotor plate. If the speed of the rotor plate is equal to that of the magnetic field, latter would be stationary when viewed from rotor plate. This is corresponding to the synchronous speed of induction motor. If the rotor plate is moved faster than this speed, the direction of the force would be reversed and a form of regenerative braking based on the principle of induction generator will come into being.

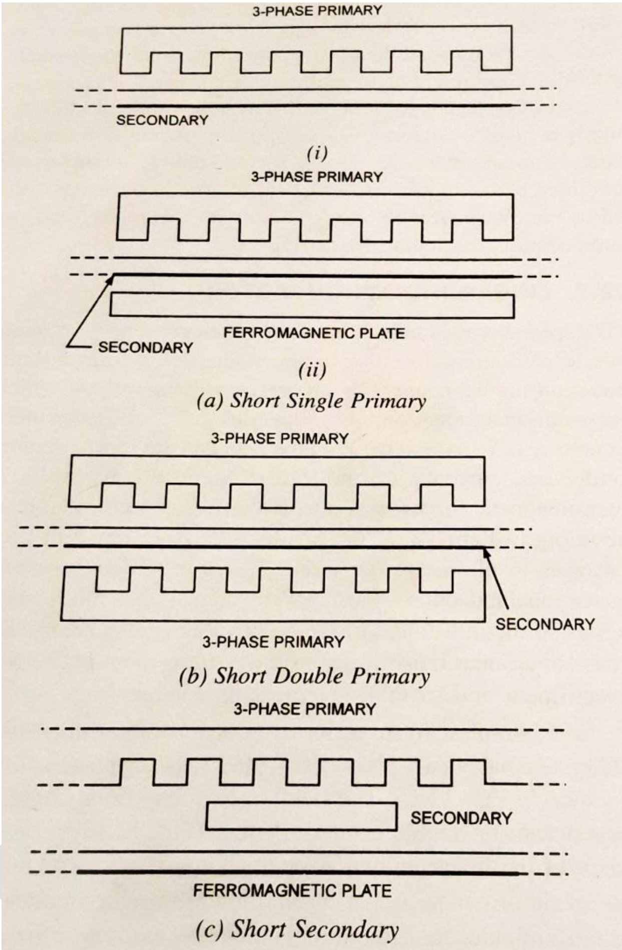


Fig. 3.8

Slip of a linear induction motor is given as

$$s = \frac{v_s - v}{v_s}$$

where v is the actual speed of rotor plate.

Thrust or force or tractive effort is given as

$$F = \frac{P_2}{v_s}$$

where P_2 is the actual power supplied to the rotor.

Active power flow is similar to that in a conventional rotary induction motor i.e.,

Copper losses in rotor = sP_2

and Mechanical power developed, $P_{mech} = (1 - s)\, P_2$

Tractive effort will be a function of slip s *i.e.,* $(v_s - v)$. The thrust (or tractive effort)-speed characteristics of a linear induction motor, as shown in figure 3.9, are similar to the torque- speed characteristics of a conventional rotary induction motor.

Tractive effort, F can be controlled by varying both frequency and Voltage simultaneously so that induction density remains constant.

There are two peculiar effects, which are encountered in a linear induction motor but not in a conventional rotary induction motor. These effects are *transverse edge effect* and *end effect.*

The paths of the induced currents in the secondary are not well defined because the secondary of a linear induction motor is a solid conducting plate. The portion of the current paths parallel to the direction of motion of secondary does not make any contribution towards the production of useful thrust but only contributes towards losses. This effect reduces the effective thrust and increases the losses and is known as the *transverse* edge *effect* because the current paths parallel to the direction of motion are more towards the edges of the conducting plate.

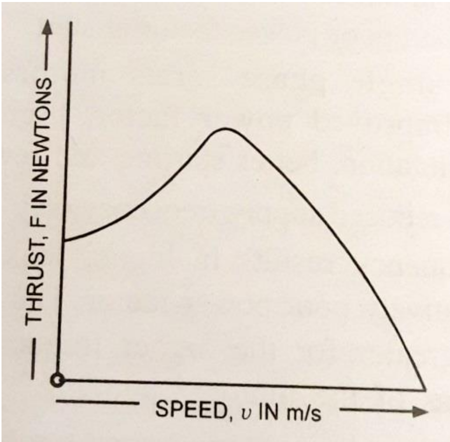


Fig. 3.9

The paths of the induced currents in the secondary are not well defined because the secondary of a linear induction motor is a solid conducting plate. The portion of the current paths parallel to the direction of motion of secondary does not make any contribution towards the production of useful thrust but only contributes towards losses. This effect reduces the effective thrust and increases the losses and is known as the *transverse edge effect* because the current paths parallel to the direction of motion are more towards the edges of the conducting plate.

In the case of linear induction motors with short primary, the current paths towards the end of field structure on the conducting plate go beyond the field structure and such portions of current-paths do not contribute to useful thrust but only towards motor losses. This is called the *end effect.* The end effect can be effectively reduced by increasing the number of poles on the motor.

The **advantages** of linear induction motors are (*i*) low initial cost *(ii)* low maintenance cost because of absence of rotating parts *(iii)* simplicity (*iv)* no limitation of tractive effort due to adhesion between the wheel and the rail (*v*) no limitation of maximum speed due to centrifugal forces *(vi)* no overheating of rotor because the motor moves continuously over cool rotor plate leaving behind heated rotor portion and (*vii*) better power to weight ratio.

The **disadvantages** of linear induction motors are (*i*) poor utilisation of motor due to transverse edge effect and end effect *(ii)* larger air gap and non-magnetic reaction rail (rotor plate) need more magnetizing current resulting in poor efficiency and low pf (*iii*) very high capital cost of reaction rail fixed along the centre line of the track (*iv*) complications and high cost involved in providing three phase collector system along the track (*v*) difficulties encountered in maintaining adequate clearances at points and crossings.

Applications: Linear induction motors can be used in conveyors, travelling cranes, haulers, electromagnetic pumps, high speed rail traction etc. Such motors may be employed in applications in which the field system moves and the conducting plate remains stationary such as travelling crane motors. These motors can also be used in applications where the field system remains stationary and the conducting plate moves such as in automatic sliding doors in electric trains, metallic belt conveyors etc. It can be used on trolley cars for internal transport in workshop, as booster accelerator for moving heavy trains from rest or up the inclines or on curves or as a propulsion unit in marshalling yards in place of shunting locomotives. Linear induction motor has superiority over conventional rotary motor for speeds over 200 kmph. Linear induction motor provides excellent source of motive power for magnetically suspended trains where conventional rotary motor fails because the latter depends for torque conversion to linear tractive force upon the adhesive weight on driving wheels. The use of linear induction motor is limited to only a few applications till now owing to design difficulties and economic considerations.

BRAKING:

Electrical and mechanical, both types of braking are used in electric traction. In electric braking the braking energy is converted into electrical energy instead of converting it into heat energy at the break shoes and either dissipated in the resistances mounted on the vehicle or returned to the supply system. Electric braking reduces the wear of the brake shoes and wheel tyres considerably and gives higher rate of braking retardation, thus brings the vehicle quickly to rest and shortens the running time to a considerable extent. In case regenerative braking is employed, the braking energy is returned to the supply system thereby a considerable saving is affected in the running cost; higher speeds are possible while going down the gradients as more braking power is available and heavier trains can be propelled down the gradients with safety and speed without dividing it into sections. Electric braking cannot replace the ordinary mechanical brakes as the vehicle cannot be held stationary by it.

The desirable requirements of a braking system are given below:

1. The braking system should be simple, robust, quick, and reliable in action.
2. Maintenance needs should be minimum and braking system must be easy for driver to control and operate.
3. The system should apply brakes simultaneously over all the vehicles.
4. Normal service application of brakes should be very gradual and smooth so as to avoid damage to the goods and discomfort to the passengers.
5. The braking force applied to each axle should be proportional to the axle load so as to obtain uniform deceleration.
6. In case of emergency braking, safety consideration is the main consideration. As such retardation rate would be maximum consistent with the safety, so as to make unfailing halt in the minimum possible distance.

7. Kinetic energy of the train should as far as possible be storable during braking which could subsequently be used during acceleration of the train.
8. The braking system should be inexhaustible.
9. There should be automatic slack adjustment for constant piston stroke as a result of wear on the rim and the brake blocks in the case of mechanical braking.

The requirements of a braking system on a main line locomotive differ from those of motor coaches. The former usually needs braking to hold the train at the steady speed on a long down grade, whereas with the latter braking is primarily required for stopping the train. Rheostatic braking is employed in both the cases but regenerative braking is mainly confined to main line locomotives due to the complications involved in providing regenerative braking with dc series motors.

There are three methods of applying electric braking namely (*i*) plugging (*ii*) rheostatic braking and (*iii*) regenerative braking.

Plugging Counter-Current Braking:

This is the simplest type of braking. In this method of braking the torque of the motor is reversed, which brings the motor and its driven machine to standstill.

In a dc motor a reversed torque is obtained by reversing the current either in the armature or in the field (not both). Polarity reversal of field winding is rarely used because it results in longer braking time due to relatively large inductance of the field winding in comparison to that can be obtained by polarity reversal of armature winding.

The connections for dc series and shunt motors during normal running and: braking conditions are shown in figures 3.10 and 3.11 respectively. In case of a dc series motor it should be ensured that the direction of flow of current in the field winding remains unchanged when the current flow in the armature winding is reversed.

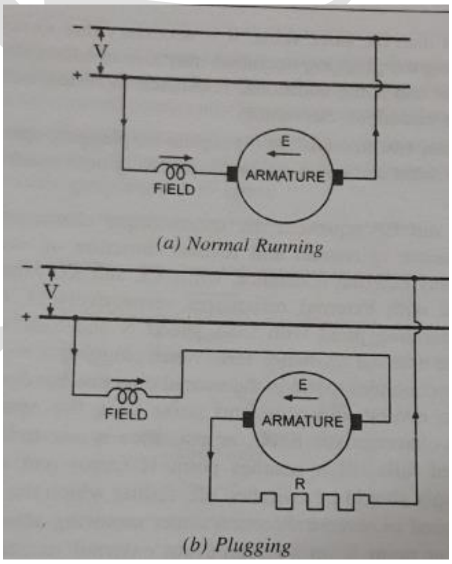
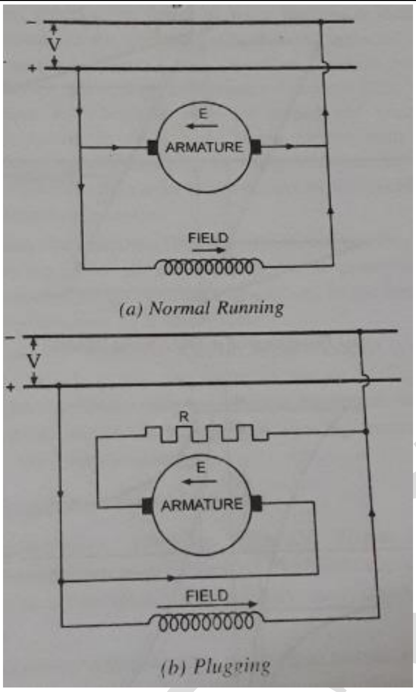


Fig. 3.10



(a) Normal Running

(b) Plugging

Fig. 3.11

In the normal running position the back emf is nearly equal to applied voltage and opposite in direction, so that a small voltage acts across the armature circuit to drive the normal current through a small resistance of the motor. In the plugging position, the induced emf acts in the direction of applied voltage, therefore, at the instant of switching the motor to the plugging position, twice of supply voltage acts across the machine circuit and heavy current would flow (about twice the current drawn by the stationary motor on normal rated voltage). Hence to avoid flow of heavy current and limit it to the safer value, it is necessary that switching performing the plugging operation may also re-insert starting resistance and some additional resistance in series with the armature circuit of the motor.

Figures 3.12 (a) and 3.12 (b) explain the plugging operation of the dc shunt and series motors respectively on a quadrantal diagram.

AB and EF represent the speed-torque characteristics of the motor in normal and reverse direction of rotation without any external resistance, while CL and KG represent the same with external resistances respectively. A is the initial operating point with load, speed N and load torque T_L and the external resistance zero. When plugging is resorted, the motor continues to run in the normal direction but develops torque in reverse direction and point G is the operating point (on characteristic KHG, as resistance is inserted here). The speed falls till it reaches point H (zero) and at this stage supply should be switched off, failing which the motor attains speed in reverse direction under motoring action and operates at point E on switching out external resistance.

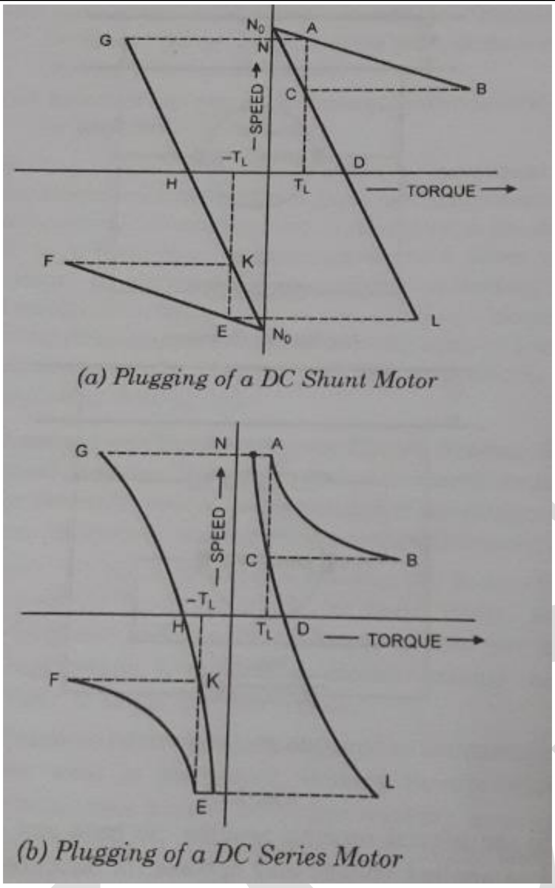


Fig. 3.12 Quadrantal Diagrams

Rheostatic or Dynamic Braking:

In this method of braking, the motor is disconnected from the supply and operated as a generator driven by the kinetic energy of the minting pails of the motor and its driven machines, thus the kinetic energy of mini ion is converted into electrical energy, which is dissipated in the external resistance connected across the motor at the braking instant.

In traction work where two or more motors are employed, they are put in parallel across a resistance for braking, as shown in figure 3.13, as the series connection would produce too high voltage. During the braking period the motors are driven as generators owing to the kinetic energy of the train and the electrical energy so generated is dissipated in the form of heat in the resistances connected across them. An equaliser connection, as shown in Fig. 3.13 (a), is used in order to ensure that the two machines share the load equally. If the equaliser connection were not done, the machine that would build up first would send a large current through the other in the opposite direction, causing it to excite with reversed voltage and so the two machines would be short circuited on themselves. *A* large braking torque would be developed in this way, but it would be quite uncontrollable and the currents would be dangerously large. An alternative method of avoiding this situation is cross-connection of the two machines, as illustrated in Fig. 3.13 (b).

In case the voltage of one of the machines becomes larger than that of the other, the first machine will cause flow of a higher current through the field winding of the second machine. As a result the second machine will have higher voltage while the field current in the first machine due to second one is smaller and hence it will cause lower voltage across first machine. Thus, an automatic compensation o/ 'the unbalancing is achieved.

The second method is advantageous to the first one as if the direction of rotation of the machine armatures reverses (may be on account of run back due to upgradient), the machines will fail to excite with equaliser connection and, therefore, no braking effect will be produced which may prove fatal for the passengers whereas with cross-connection the machines will build up in series and being short circuited on themselves will provide an emergency braking and the train will not be allowed to run back on the gradient. ***Rheostatic braking cannot be employed with 3-phase induction motors.***

In case of ac series motors the rheostatic braking is obtained by operating the machines as generators excited from the supply or as self-excited dc generators supplying power to resistance load. In the former case the fields are energised at low voltage from a suitable tapping on the main transformer while in the latter case the fields of the motors are excited from one of the motors acting as a series generator and in this case dc will be generated in rotors of the motors and the kinetic energy of the rotors will be dissipated as dc power in the loading resistors.

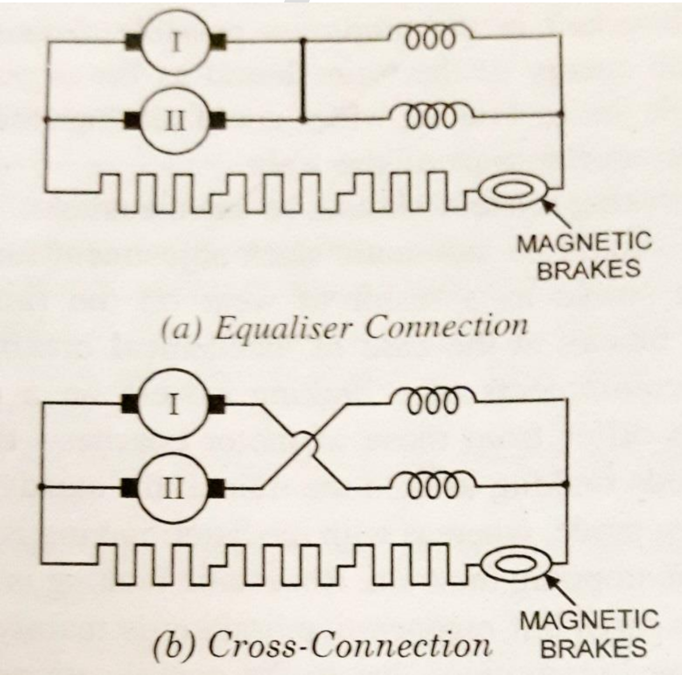
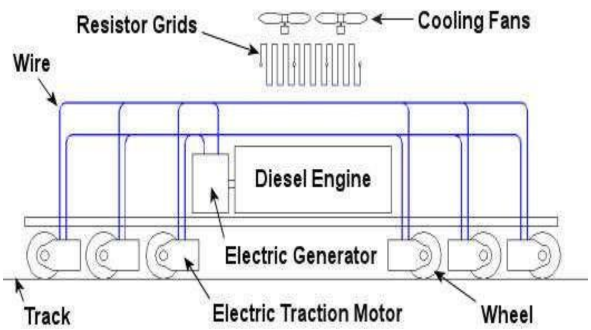


Fig. 3.13

The generated energy is sent to the resistor grids in the roof of the train and is dissipated as heat.



Regenerative Braking:

In the previous two methods of electric braking namely plugging and rheostatic braking, stored energy of the rotating parts of the motor and its driven machines is wasted whilst in plugging extra energy is drawn during the braking period and wasted. In the regenerative braking the motors remain connected to the supply and return the braking energy to the supply.

For regenerative braking it is necessary that traction motors I must generate power at a voltage higher than the supply voltage and at a reasonable constant voltage.

Regenerative braking is an inherent characteristic of *shunt wound motors* and does not require any change of connections. DC shunt motors are electrically stable also *i.e.*, braking torque is independent of line voltage fluctuations.

Regenerative Braking With DC Shunt Motors:

Regenerative braking is an inherent characteristic of shunt wound motors and does not require any changing of connections.

Let us consider a shunt motor running in normal condition. Now if due to, say, hauling load, speed increases above normal, due to increase in speed the induced emf will increase and , if it exceeds the line voltage, the machine will start supplying current to the line, thus have tendency to prevent any further increase in speed.

Similarly if the field current is increased so that motor emf exceeds the line voltage and it may start supplying current to the line, the motor will slow down to the speed corresponding to this new value of field current.

Regenerative braking can be easily applied to dc shunt motors, particularly in cases where it is required to hold a load at a certain speed for instance lowering a hoist, it is not however, possible to obtain regenerative braking down to very low speeds, since a sufficient increase in field current cannot be obtained.

Before bringing the motor to rest, the field excitation is increased to the permissible maximum value, as a result of which the speed of the motor is reduced to the minimum value and the kinetic energy released from the motor is fed back to the supply.

Since the motor is required to operate with a weak field at rated condition, the armature has to be designed to carry a large current to produce the rated torque and hence motor will be larger in size, poor in efficiency and costlier.

DC Shunt Motor in Hoisting Mechanism: Consider a dc shunt motor (or a separately excited motor) used in hoisting mechanism. When it is switched on for lowering a load, the torque developed by the motor and that due to load torque act in unison and, therefore, accelerates the motor. With the increase in speed, induced emf in the armature (i.e., back emf, E) increases till E = applied voltage V. At this moment the speed becomes the ideal no-load speed, the armature current and, therefore, the developed electromagnetic torque is zero, so that the downward motion of the hoist is sustained only by the downward moving load. When the speed exceeds the ideal no-load speed, induced emf E exceeds the applied voltage V and, therefore, armature current becomes negative. The drive, then acts as a generator and provides the braking torque. Under such condition

$$\text{Current } I = \frac{E - V}{R}$$

and electric braking torque,

$$T = K_1\phi I = K_1\phi\ \frac{E - V}{R}$$

$$= K_1\phi\ \frac{K_2 N\phi\ - V}{R}$$

where K_1 and K_2 are constants.

or braking torque,

$$T = \frac{K_1 K_2 \phi^2 N}{R} - \frac{K_1\phi V}{R}$$

$$= K_3\phi^2 N - K_4\phi$$

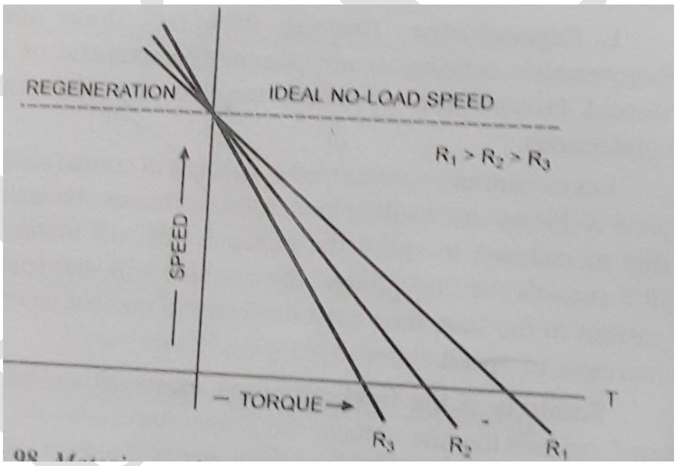


Fig. 3.14 Motoring and Regenerative braking characteristics of a DC Shunt Motor

The speed-torque characteristics of motor during motoring and braking operation are shown in figure 3.14. From figure 3.14 it is obvious that braking or regeneration can occur only speeds greater than the ideal no-load speed of the motor. From figure 3.14 it can also be noted that higher the resistance of the armature circuit, to develop a particular braking higher is the speed required.

Regenerative Braking With DC Series Motor:

The dc series motors cannot be used for regenerative braking in an ordinary way. Since the reversal of armature current necessary to produce regeneration would cause a reversal of field, therefore, series field connections must be reversed. But even if the field connections are reversed at the exact moment, this method would still be useless. Because at the instant of reversal, the emf induced in the motor will be small, so current will flow through the field in wrong direction, which will reverse the field and cause the motor emf to help the supply voltage. This will result in short circuit of supply. Due to these complications this method is not used for common industrial purposes. Regenerative braking is, however, used with series motors for traction either by modification of windings or by supplying the machines with separate excitation.

One method of obtaining regenerative braking with series motors is the French method. If there is a single series motor as in case of a trolley buses, tramways, it is provided with a main series field winding and auxiliary field windings connected in parallel with the main series field winding shown in figure 3.15 (a). During regeneration (braking period) the auxiliary field windings are put in series with each other and are switched across the supply, as shown in figure 3.15(b). The machine acts as a compound generator slightly differentially compounded. Such an arrangement is quite stable. Any change in line voltage causes a change in excitation which produces a corresponding change in the induced emf of the machine so that inherent compensation is provided. For example, if the line voltage increases beyond the emf of the generator the increased voltage across the generator's field will send a large exciting current through it causing the emf of the generator to rise. The reversal of this will happen when the line voltage decreases.

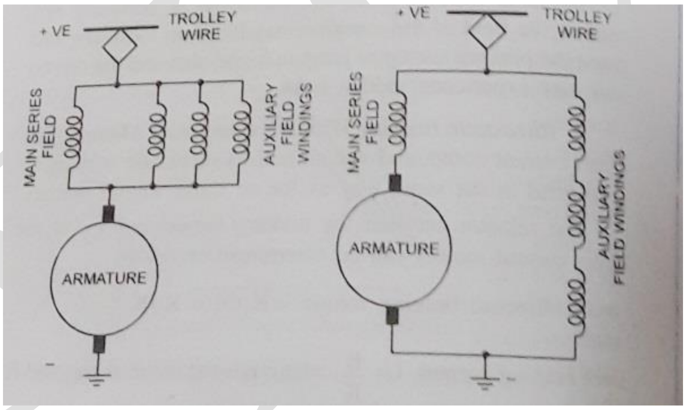


Fig. 3.15 French Method of Regenerative Braking

If there are several motors, we do not require any auxiliary winding. During normal running the motors are connected in parallel with the field winding connected in series with their respective armatures, as shown in figure 3.16(a).

But during regeneration the motors are connected, as shown in figure 3.16 (b), *i.e.*, all armatures are connected in parallel and series field windings of all motors but one are connected in series and placed across the supply. Suitable resistance is also connected in series with the series field windings, as illustrated in the figure 3.16 (b).

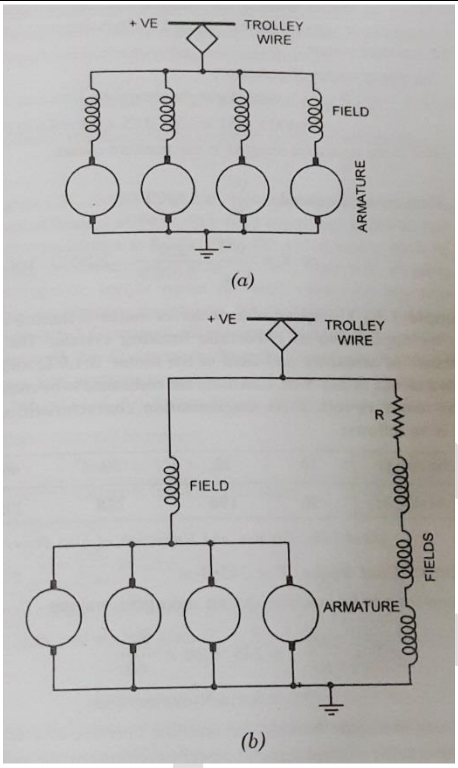


Fig. 3.16

Another alternative method of regenerative braking, which obviates the loss of power in the resistances in the Held circuit, is by using a separate exciter for controlling the excitation of the field windings during regeneration. The exciter may either be axle driven or driven by a motor operated from the auxiliary supply. Many devices have been used to secure compensation against variations in line voltage, most of which make use of some from of differential winding on the exciter. One such method is illustrated in figure 3.17 *(a)* in which the exciter is differentially compounded. The shunt field winding of the exciter is separately excited from an auxiliary supply; the series field winding is connected in the main-motor circuit in such a way that it opposes the separately excited winding during regeneration. It can be seen that a rise in line voltage during regeneration, which will tend to reduce the regenerated current, will strengthen the exciter field and counteract the change. The stabilising resistance R also assists in this action.

Another method used for regenerative braking is depicted in figure 3.17 *(b)*. Here the exciter armature along with the field windings of the series motors are connected across the stabilizing resistance in the main circuit. The current through the stabilizing resistance is the sum of the exciter current and the regenerated current Any increase in regenerated current due to fall in line voltage, causes larger voltage drop in the stabilizing resistance and, therefore, less current through the exciter circuit again counter-acting the effect of line voltage variations.

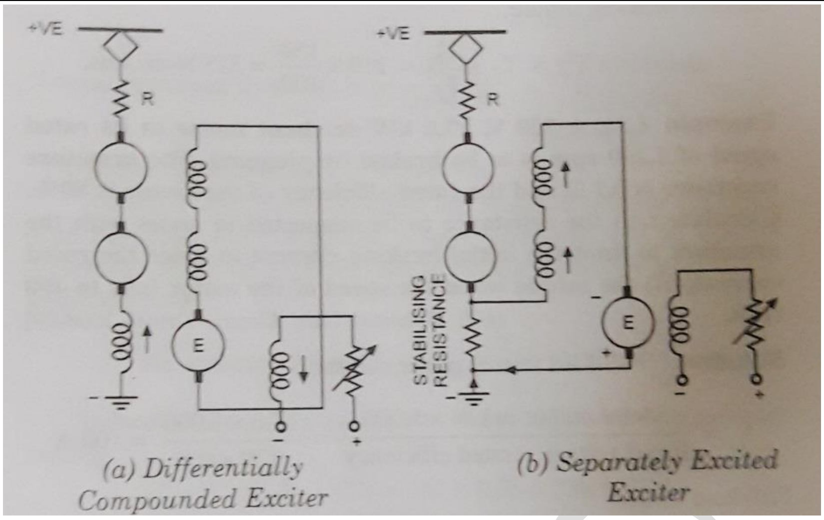
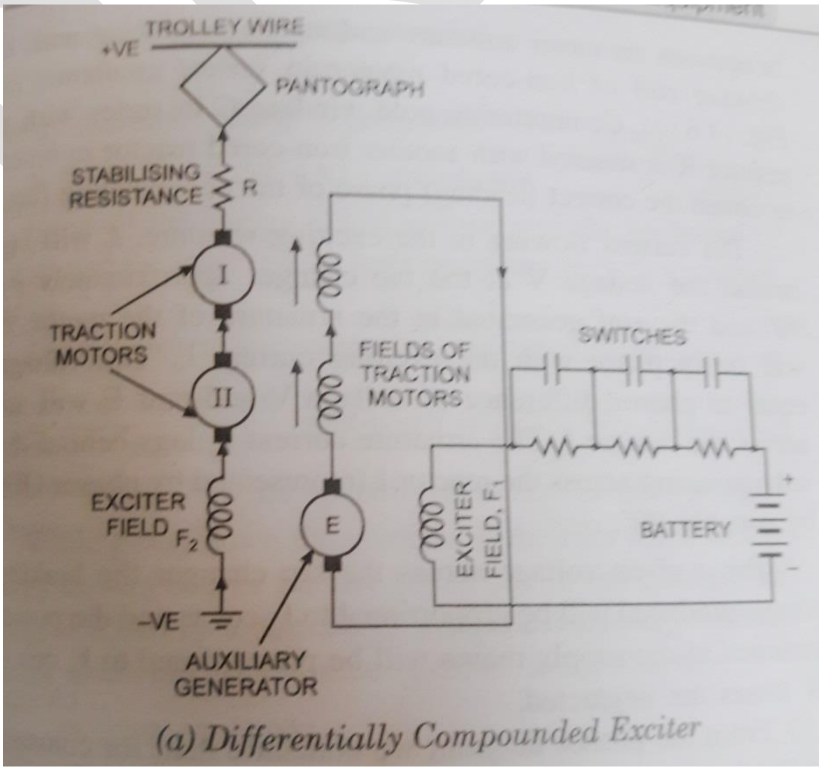
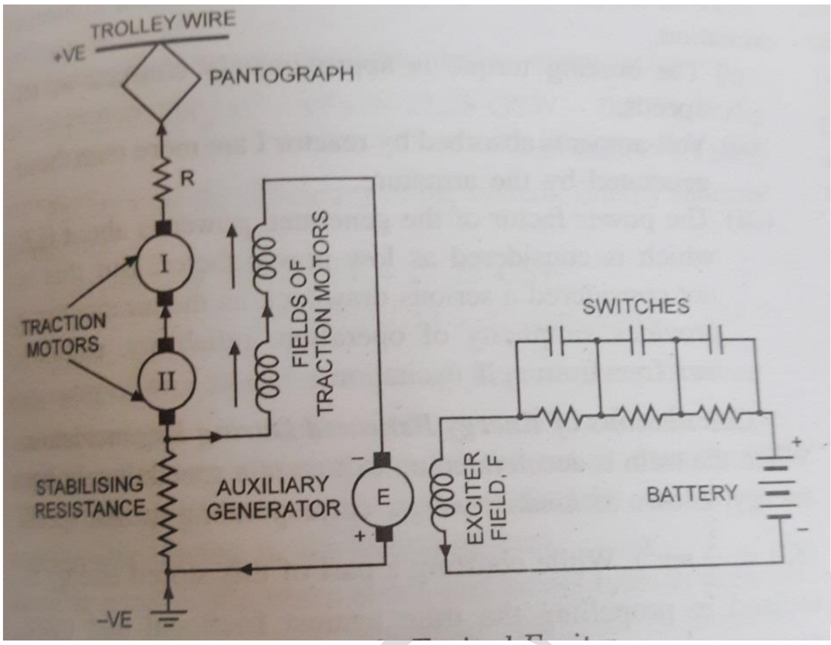


Fig. 3.17. Regenerative Braking With Separate Exciter

Inherently a dc series generator has no stability of voltage, and therefore, an external means is required. In figure 3.18 (a) a separate motor driven exciter has a separately excited field winding F_1, and also another field winding F_2 connected in the main motor circuit in such a fashion that the field created by it opposes the field created by separately excited field winding F_1 during regeneration. The stabilizing resistance R is employed to prevent current surges when the tram crosses from one section of the supply to another, and to compensate for variable line voltage. In case the line voltage falls, regenerated current will tend to increase resulting in strengthening of field F_2 which will weaken the field F_1, and, therefore, reduction in emf generated by the exciter. Thus the field of the traction motors will be weakened resulting in reduction of emfs generated by the traction motors operating as dc series generators. Thus compensation for a decrease in the line voltage is automatically provided.





(b) Separately Excited Exciter

Fig. 3.18 Regenerative Braking With Separate Exciter

The arrangement shown in figure 3.18 (b) has the exciter armature connected in the circuit of the field windings of traction motors and the stabilising resistance. The current through the stabilising resistance is the sum of the exciter current and the regenerated current. The voltage of the exciter circuit can be regulated by either varying its field strength or manipulating resistances in series with the armature. In case the line voltage falls the regenerated current will tend to increase resulting in increase in voltage drop across the stabilising resistance and, therefore, reduction in the voltage available in the exciter armature circuit causing a reduction in the excitation current of the traction motors operating as dc series generators. This reduces the emf's generated and thus compensation is provided automatically.

Regenerative braking is an inherent characteristic of an *induction motor* since an induction motor operates as an induction (non-synchronous) generator when run at speeds above synchronous speed and it feeds power back to the supply line. The machine, however, is not self-exciting as a generator and is required to be connected to a system supplied from synchronous generators. This system supplies the excitation and determines the frequency at which the induction generator operates. Torque-speed characteristics of an induction machine are shown in figure 3.19. From the torque-speed characteristics of an induction machine it is obvious that without any extra resistance in the rotor circuit, there is only a slight variation of speed with the torque whereas with the external resistance inserted in the rotor circuit speed increases for a particular braking torque. Thus we see that with no extra resistance in the rotor circuit the speed during braking remains almost constant and independent of the gradient and the weight of the train. This is a great advantage with the induction motor when used for traction. But if increased speeds are necessary with light loads, these can be obtained by inserting external resistance in the rotor circuit. It is advantageous on mountain railways. It returns about 20% of the total energy on certain railway run and saves a great deal of brake shoe wear.

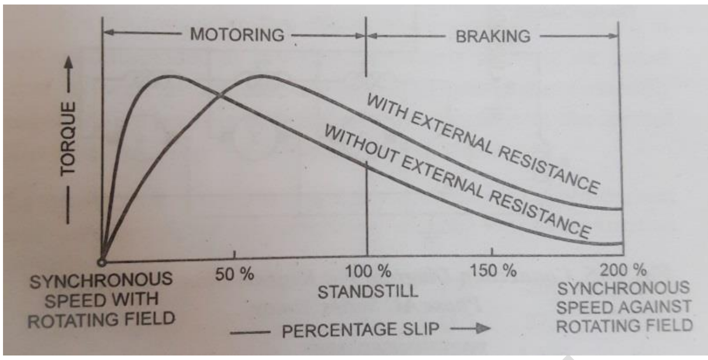
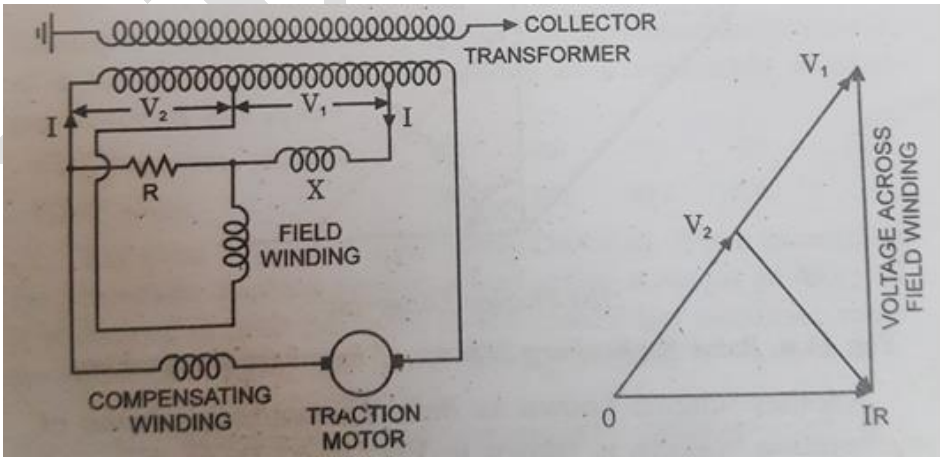


Fig 3.19

Regenerative braking with ac series motors is more difficult than that with the dc series motor. The difficulties encountered are as follows.

During the regeneration period, the machine should not operate as a self-excited generator. Generally the high power factor is the main consideration in order to achieve a reasonable braking torque. Circuit for regeneration at high power factor usually suffers from self-excitation, and those which are inherently stable and free from self-excitation usually operate at a poor power factor.

For regenerative braking the regenerated power should have the same frequency as that of the main supply. This necessitates the energizing of the motor field winding from the ac mains. The regenerated current must be in phase opposition to the applied voltage and also the flux ϕ, so that power may be supplied back into the supply system. The arrangement to provide the above conditions is illustrated in figure 3.20.



(a) Connection Diagram (b) Phasor Diagram

Fig. 3.20 Regenerative Braking With Single Phase AC Series Motor

In another satisfactory arrangement one of the traction motors is used as a generator to supply current to excite the fields of the remaining motors. As the speed of the locomotive falls, the voltage of the motors can be controlled by increasing the excitation current. This is accomplished by increasing the voltage on the exciting motor by employing another transformer tap. Further flexibility can be obtained by changing the tap from the generating motor (Fig. 3.21).

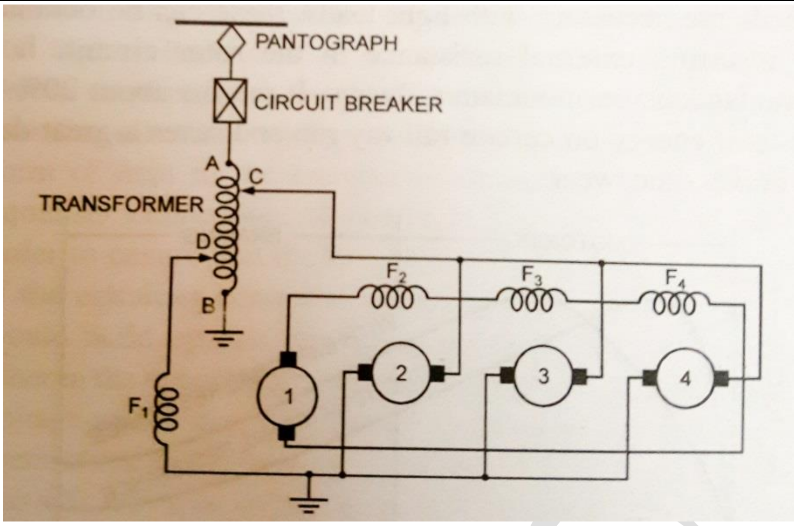
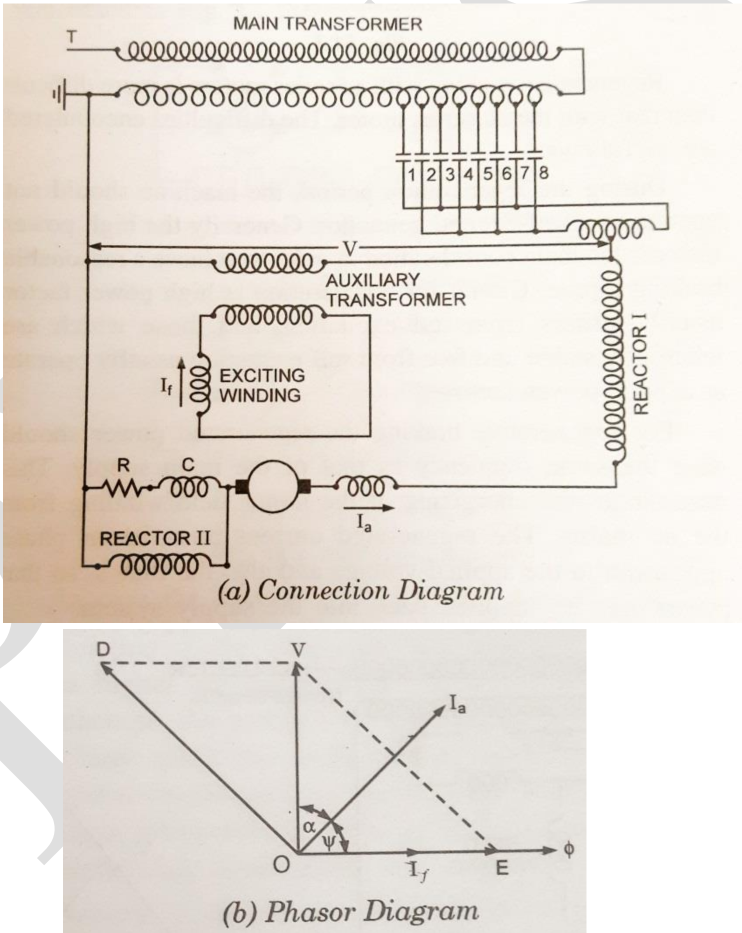


Fig. 3.21. Connection Diagram for Regeneration With Single Phase AC Series Motor



(a) Connection Diagram

(b) Phasor Diagram

Fig. 3.22. Behn Eschenburg Scheme of Regenerative Braking

Another scheme known as *Behn Eschenburg scheme* of regenerative braking is shown in figure 3.22 (a). In this case an auxiliary transformer is used to excite the exciting winding of the traction motor. The armature of the traction motor is connected to the main transformer through tap changer.

In between the motor armature and tap changer is inserted a choking coil or iron-cored reactor in series, as shown in figure 3.22 (a). Commutating pole winding C in series with a resistor R is shunted with another iron-cored reactor in order to obtain the correct (leading) phase of the commutating flux.

The current flowing in the exciting winding, I_f will lag behind the voltage V at the tap changer approximately by 90° and the emf generated in the armature of the motor, E will be in phase with the exciting current I_f. The voltage equal to phasor difference of voltage V and emf E will act across the reactor I. The armature current I_a lags behind the voltage acting across the reactor I (represented by phasor OD) by roughly 90°.

For a given voltage across the tap changer the braking torque produced will be proportional to $I_a \cos \psi$ and the power returned to the supply mains will be proportional to $I_a \cos\alpha$ if losses are neglected.

From the phasor diagram we conclude that, for constant excitation,

(i) The braking torque is approximately constant at all speeds.
(ii) Volt-amperes absorbed by reactor I are more than those generated by the armature.
(iii) The power factor of the generated power is about 0.7, which is considered as low power factor. But this is not considered a serious drawback as the arrangement provides simplicity of operation, reliability, stability and free from self-excitation.

Regenerative Braking with DC Compound Motors:

Compound wound motors can be regeneratively retarded by a hauling load only if their shunt field is stronger than their series field.

Regenerative Braking of three phase induction motors:

Regenerative braking is an inherent characteristic of an induction motor, since it operates as an induction generator when it runs at speed above synchronous and it feeds power back to the supply line.

The 3-phase induction motor can be made to operate at speed above synchronous speed by employing any one of the following processes.

i. Switching over to a low frequency supply in frequency controlled induction motors in order to reduce the speed of operation of the drive.
ii. Downward motion of a loaded hoisting mechanism such as cranes, hoists, excavators etc.
iii. Switching over to a larger pole number operation from a smaller one in multi-speed squirrel cage motors.

In all the above processes, the slip and torque developed become negative, as shown in figure 3.23, and thus the machine acts as a generator, receiving mechanical energy and giving it back to the supply system in the form of electrical energy.

If the load drives the motor above synchronous speed, no switching operation is required. Once the machine is driven above synchronous speed, the braking operation automatically starts. The operating point will depend upon the magnitude of load torque and the nature of torque-speed characteristic of the machine during generating operation. By varying the resistance in the rotor circuit, it is possible to operate at any speed above synchronous speed during braking. In case the driving torque of the load exceeds the maximum braking torque, of which the machine is capable, the system will become unstable and the speed will rise further, probably to a disastrous value, since, the faster the machine runs, the lesser will be the braking torque developed.

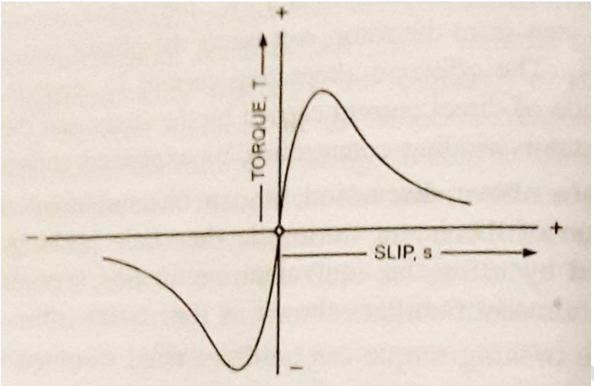


Fig. 3.23

In the case of a squirrel cage induction motor, stable speed is obtained at a speed considerably in excess of the synchronous speed and the regenerative braking cannot be applied unless the motor is specially designed to withstand the excessive speed.

Regenerative braking has the disadvantage of the possibility of braking only at super-synchronous speeds and, therefore, is seldom used for braking. This method can be used only in hoisting type of mechanism or with a multi-speed squirrel cage motor. It is advantageous on mountain railways too. It returns about 20% of the total energy on certain railway runs and saves a great deal of brake shoe wear.

Advantages of Regenerative Braking:

1. A part of energy is returned to the supply system, so energy consumption for the run is considerably (about 20 to 30 per cent) reduced thereby affecting a considerable saving in the operating cost.
2. The wear of the brake shoes and wheel tyres is reduced to considerable extent, therefore, their life is increased and replacement cost is reduced.
3. Higher value of braking retardation is obtained so that the vehicle can be brought to rest quickly and running time is considerably reduced.
4. Small amount of brake dust is produced when the mechanical brakes are applied.
5. Higher speeds are possible while going down the gradients because the high braking retardation can be obtained with regenerative braking.
6. Propulsion of heavier trains on gradients is possible without dividing them into sections with speed and safety.

Disadvantages of Regenerative Braking:

1. Additional equipment is required for control of regeneration and for protection of equipment and machines, hence initial as well as maintenance cost is increased.
2. The dc machines required in case of regenerative braking are of large size and cost more than those ordinarily employed, therefore, the weight of the locomotive and thus the required mechanical strength and cost increase.
3. Owing to recuperated energy the operation of the substations becomes complicated and

difficult.

4. In case of substations employing mercury-arc rectifiers for conversion purpose, additional equipment is required either to deal with regenerated energy separately or to change one or more of the ordinary rectifiers over to inverted operation. No such difficulty is experienced in case of substations employing rotary convertors or motor-generator sets for converting purpose.
5. Regenerative braking is employed down to a speed of 16 kmph, then rheostatic braking to about 6.5 kmph and then mechanical braking is required to bring the locomotive to rest.

In most of the cases, however, and especially with motor- coach trains the increased cost of the train equipment and the additional features required in order to obtain regenerative braking, combined with the increase in the maintenance cost of the electrical equipment, may entirely offset the economics in the energy consumption and the other items.

For tramways, trolley buses, the regenerative braking is not recommended as it will unnecessarily increase the initial cost as well as increase the operating problems. Generally regenerative braking is desirable and necessary for the service lines having long gradients exceeding 0.6%.

Calculations of Energy Returned During Regeneration:

When the train is accelerated up to a certain speed, it acquires energy, known as *kinetic energy***,** corresponding to that speed $\left(\mathrm{KE} = \frac{1}{2} m v^2\right)$. While coasting a part of this stored energy is utilized in propelling the train against frictional and other resistances to motion and, therefore, the speed falls. Under ideal conditions (no resistance to motion of the train) the speed of the train would have not decreased.

Similarly while the train going down the gradient or moving on level track, the speed remaining the same or reduced, this stored energy can be converted into electrical energy and returned back to the lines.

The amount of energy returned to the line depends upon (*i*) the initial and final speeds during regenerative braking **(***ii*) the train resistance and gradient of the track also in case the train is moving down the gradient and **(***iii***)** efficiency of the system.

Let the initial and final speeds of the train be V_1 and V_2 kmph respectively. The kinetic energy stored in the train at a speed of V_1 kmph

$$= \frac{1}{2} \times \frac{1{,}000 M_e}{9.81} \left(\frac{1{,}000 V_1}{3{,}600}\right)^2 \quad \mathrm{kg-m}$$

$$= \frac{1}{2} \times 1{,}000 M_e \left(\frac{1{,}000 V_1}{3{,}600}\right)^2 \quad \mathrm{N-m\ \ or\ \ Watt-seconds}$$

$$= \frac{1}{2} \times 1{,}000 M_e \left(\frac{1{,}000 V_1}{3{,}600}\right)^2 \times \frac{1}{3{,}600} \quad \mathrm{N-m\ \ or\ \ Watt-hours}$$

$$= 0.01072\ M_e V_1^2 \ \ \mathrm{Watt-hours}$$

Similarly kinetic energy at speed V_2 kmph

$$= 0.01072\ M_e V_2^2 \ \ \mathrm{Watt-hours}$$

Energy available during regeneration

$$= 0.01072\ M_e(V_1^2 - V_2^2)\ \text{Watt} - \text{hours}$$

Some of the energy is lost to overcome the resistance to motion and the losses in the traction system including motors.

Energy lost to overcome the resistance to motion

$$= \frac{M \times r \times S \times 1{,}000}{3{,}600}\ \text{Watt} - \text{hours} = 0.2778\ \mathrm{M}rS\ \ \text{Wh}$$

where r is the specific resistance in newton/tonne and S is the distance travelled during regenerative period.

While going down the gradient in the hilly track service, energy is supplied as tractive effort due to the gradient and energy is added up to the energy available during regeneration.

Energy available due to motion down the gradient

$$= \frac{98.1GM \times S \times 1{,}000}{3{,}600} = 27.25\ GSM$$

Hence total energy available during regeneration

$$= [0.01072\ M_e(V_1^2 - V_2^2) + 27.25\ GSM - 0.2778\ \mathrm{M}rS]\quad \text{Watt} - \text{hours}$$

Taking η as the efficiency of the system, Energy returned to the line

$$= [0.01072\ M_e(V_1^2 - V_2^2) + 27.25\ GSM - 0.2778\ \mathrm{M}rS] \times \eta\quad \text{Watt} - \text{hours}$$

Average Power Calculation:

$$\text{Average speed } = \frac{V_1 + V_2}{2}\ \text{kmph}$$

$$\text{Average time taken to cover S km} = \frac{\text{S}}{\text{Average speed}}\quad \text{hours}$$

$$\text{Average power, P} = \frac{\text{Energy returned to the line}}{\text{Time}}\quad \text{Watt}$$

Train Lighting System:

Train lighting is one of the important passenger amenities which influence the image of Railways. Although first train ran on 16th April 1883 from Mumbai CST to Thane, train lighting system through axle driven dynamo pioneered by M/s. J. Stone & Co. came to Indian Railways only by 1930. Dynamo / Brushless alternator driven from axle through flat / 'V' belts, supplies the load when train is in motion and charges the batteries. The batteries supply the load when train is stationary. Following systems for train lighting are presently in use –

1) Axle driven system working on 110 V DC supply.
2) Mid on generation with 415 V, 3 Phase generation AC 110 V utilization.
3) End on generation with 3 Phase 415 V generation and AC 110 V utilization
4) End on generation with 3 Phase 750 V generation and AC 110 V utilization

AXLE GENERATION WORKING ON D.C. 110 V SUPPLY

This system has proved more reliable and capable of meeting future increase in load. It has, therefore, been adopted as standard for all future builds of self-generating, coaches. In this system 4.5 KW brushless alternators are driven through V-belts from axle. Lead acid batteries 110 V, 120 Ah arranged from 3 cell Monoblock units, are provided in the B.G. coaches. Four numbers of emergency feed terminals boxes for B.G. (Brake Gangwayed or BG coach) and one number for M.G. coach, are provided on each end wall for interconnecting the coach to adjacent coach to receive power, in the case generation fails. One number emergency terminal box is provided centrally on each side of under frame to facilitate charging of battery from external source. For BG AC coaches, 18 KW / 25 KW brushless alternators are used. Two such alternators are used in AC-2T /AC-3T /Chair Cars and only an alternator is used in First AC coach. Batteries of 800 / 11 00 AH capacity at 10 hr rating are used in I AC / AC-2T / AC-3T /chair car of B.G. Coaches. A schematic layout for 110 V DC system is at figure 3.24.

Three phase output from 4.5 kW alternator mounted on the bogie of coach is fed to the regulator cum rectifier for rectifying the AC output to DC and regulating the output voltage at different speeds and loads. The output from rectifier cum regulator on the under frame is brought through cables on the coach. The load is fed through four rotary switches (RSW) and fuses connecting circuits L_1, L_2, F and SPM. L_I feeds the essential lighting load like lavatories, gangways, doorways and upto 50% of light in each compartment/bays corridor lights and night lights, L_2 feeds remaining lighting loads, F feeds the fan load and SPM feeds emergency feed terminals (EFT).

An external battery charging terminal (BCT) is provided to charge the battery from external charger, if battery is in rundown condition due to failure of alternator.

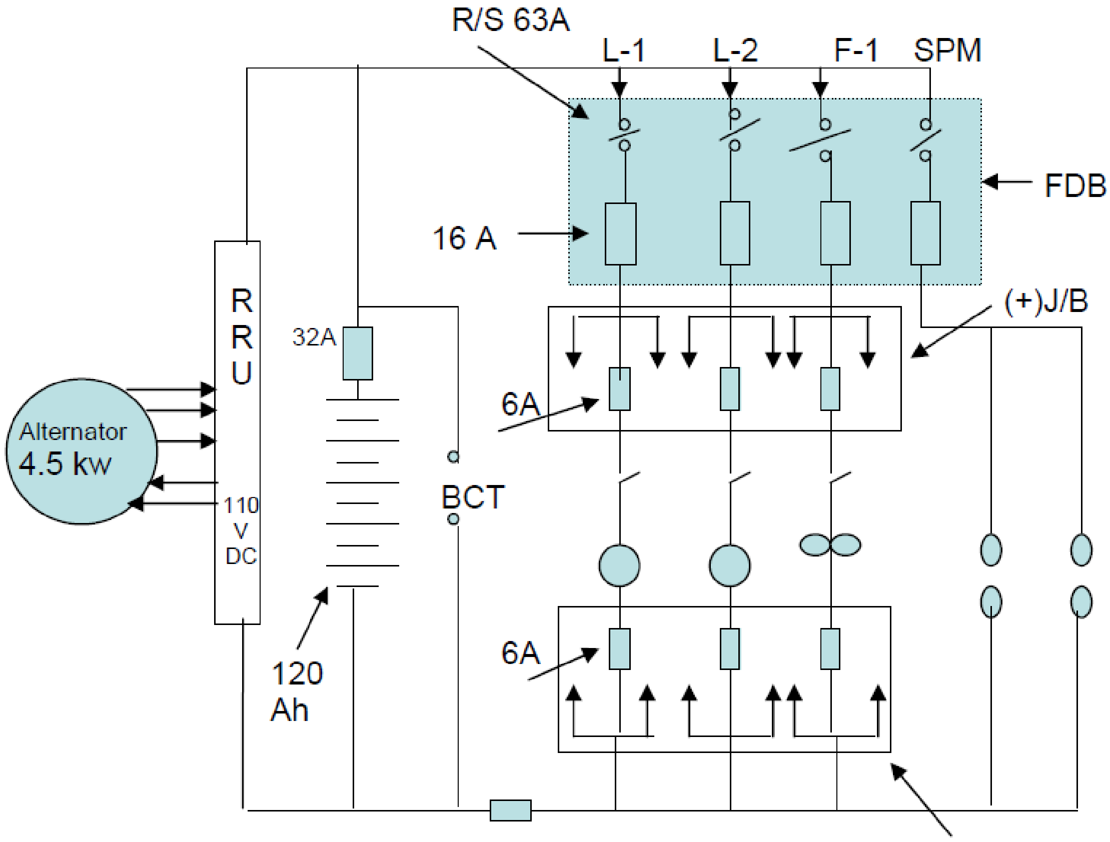


Fig. 3.24 Schematic diagram of 110V DC Train Lighting System

MID-ON-GENERATION:

In this system a power car housing DG sets is used in middle of rake. This system is chosen for small branch line slow trains having long halts where batteries are likely to remain undercharged if conventional axle driven system is adopted. Capacity of DG set will depend on composition of rake (usually 30 kVA) and generation is at 415 V, 3 phase, 50 Hz and is stepped down to 110 V, 3 Phase, 50 Hz.. The lights and fans in coaches are operated 110V AC through feeders on either side of Power Car.
A schematic layout of power car for mid-on-generation is shown at figure 3.25.

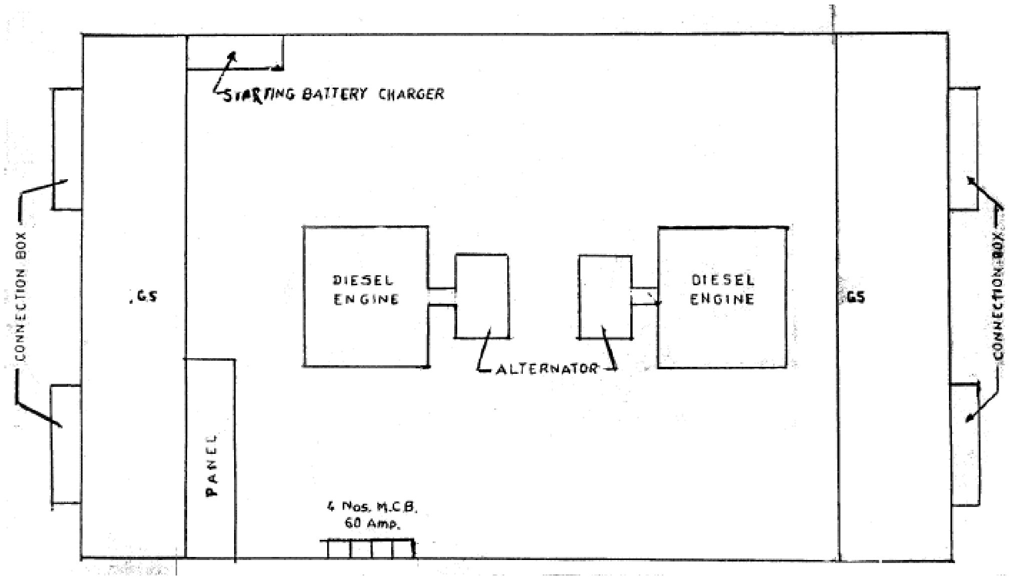


Fig. 3.25 General Layout of MOG Power Car

END-ON-GENERATION:

Rakes of Rajdhani / Shatabdi express trains having heavy load of air-conditioned coaches, pantry cars with electrically operated cooking appliances, use Diesel Generating Sets housed in coaches known as Power cars to meet the load. Normally 2 power cars, one on either side of rake, generate power at 750 V AC or 415 V AC, 3 phase, 50 Hz. All the coaches of power cars are interconnected with each other through couplers consisting of switchgear flexible cables. Power cars have control panel consisting of switchgear and protective relays, the power at 750 V/ 415 V is stepped down to 110 V AC for lighting and fan load in the coaches. A schematic layout of power car for end-on- generation is shown at figure 3.26.

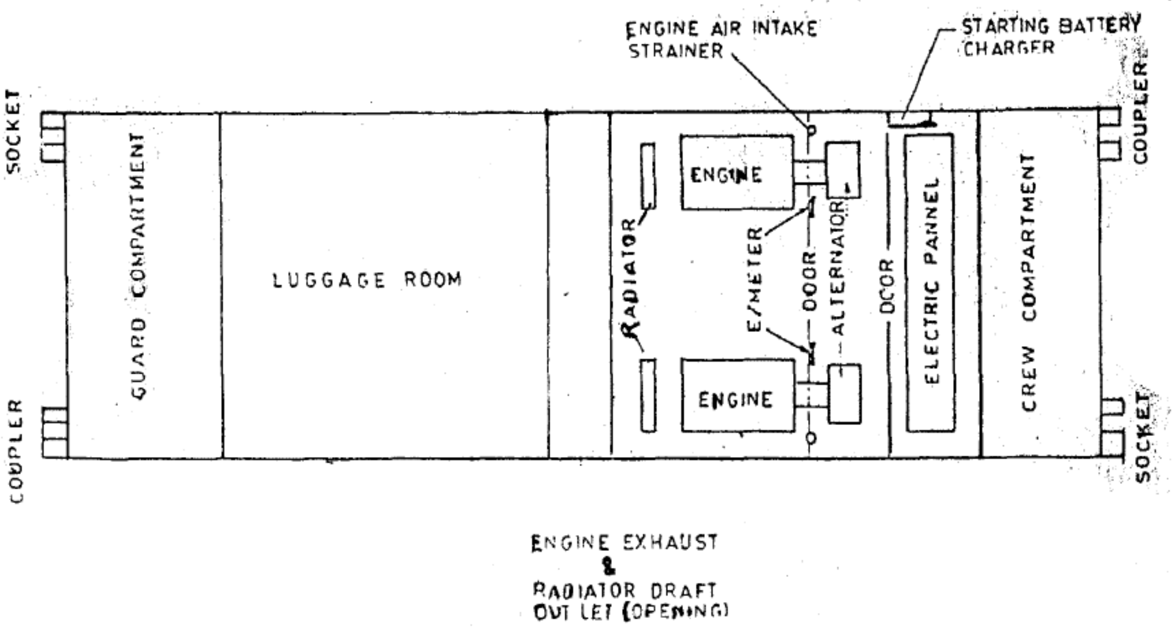


Fig. 3.26 General Layout of EOG Power Car

HEAD-ON-GENERATION:

In this system, power is provided from the locomotive at the head of the train. Electric locomotives derive power from overhead supply lines. Power produced is 4500kW. The power produced is supplied to all loads in AC, Non- AC coaches. Diesel locomotives use a separate Diesel generator sets to generate electric power. Diesel locomotives have been widely replaced by electric locomotives.

Advantages:

- Pollution free
- Cost of power production per unit is 25% cheaper than EOG & SG
- More commercial space available due to elimination of power cars.

Disadvantages:

- Power interruption of short duration

Questions Bank

1. List the equipments for overhead A.C. traction system and explain with a neat sketch pantograph collector.
2. What is regenerative braking? List the advantages of regenerative braking in series motors.
3. Describe the characteristics of AC series motor. Explain how it is suitable for traction system. With a neat figure, explain the working principle of linear induction motor.
4. Mention the advantages of serial-parallel starting method and explain what is the practical requirement of it?

5. Explain the different methods of speed control of induction motors.
6. Discuss the characteristics of A.C series motors used in electric traction.
7. Write a note on diesel-electric traction equipment.
8. What is braking and explain the regenerative braking system employed in DC shunt wound motor.
9. Explain linear induction motor.
10. Discuss the suitability of D.C. series motor for electric traction application.
11. What are the advantages of regenerative braking? Explain how regenerative braking can be obtained in dc locomotives.
12. Explain the theory and working characteristics of linear induction motor for traction.
13. Mention the different systems of traction and list out their advantages and disadvantages. Explain the series parallel control of traction motors.
14. Discuss the characteristics and the suitability of single phase ac series motor for traction operation.
15. Write notes on i) Linear IM ii) Regenerative braking.
16. Explain clearly the train lighting system and its accessories.
17. Discuss the suitability of D.C shunt and series machines for regenerative braking.
18. A train weighing 500 tonnes is going down a gradient of 20 in 1000. It is desired to maintain train speed at 40 km/hr. by regenerative braking. Calculate the power fed in the line, traction resistance is 40N/tonne and allow rotational inertia of 10% and % efficiency of conversion is 75%.
19. A 400T electric train has its speed reduced from 60 to 40 km/ph by regenerative braking over a distance of 2km along down a gradient of 1.5%. Assuming train resistance as 50Nw/T, rotational inertia 10%, efficiency of conversion 75%, calculate (i) Electrical energy returned to the line, (ii) Average power returned to the line.
20. A 400 tonne train travels down a gradient 1 in 70 for 120 sec during which period its speed is reduced from 80 km/hr to 50 km/hr by regenerative braking. Find the energy returned to the lines if the tractive resistance is 5 kg/tonne and allowance for rotational inertia is 7.5%. overall efficiency of motors is 75%.
21. A 2,340 tonne train (including loco) proceeds down a gradient of 1 in 80 for 5 minutes during which period, its speed gets reduced from 60 kmph to 36 kmph by application of regenerative braking. Find the energy returned to the lines, if the tractive resistance is 5 Kg/tonne, rotational inertia 10% and the overall efficiency of this motor during regeneration is 70%.
22. A locomotive accelerates a 350 ton train in an up-gradient of 1 in 100 at 0.8 km per hour per second. Taking the coefficient of adhesion to be 0.25, determine the minimum adhesive weight of the locomotive. Assume the train resistance of 45 Newton per tonne and allow 10% for the effect of rotational inertia.

UNIT-4

ILLUMINATION

Radiations from a Hot Body:

The usual method of producing artificial light consists in raising a solid body or vapour to incandescence by applying heat to it. It is found that as the body is gradually heated above room temperature, it begins to radiate energy in the surrounding medium in the form of electromagnetic waves of various wavelengths. The ***nature*** of this radiant energy depends on the temperature of the hot body. Thus, when the temperature is low, radiated energy is in the form of heat waves only but when a certain temperature is reached, light waves are also radiated out in addition to heat waves and the body becomes luminous. Further increase in temperature produces an increase in the amount of both kinds of radiations but the colour of light or visible radiation changes from bright red to orange, to yellow and then finally, if the temperature is high enough, to white. As temperature is increased, the wavelength of the visible radiation goes on becoming shorter. It should be noted that heat waves are identical to light waves except that they are of longer wavelength and hence produce no impression on the retina. Obviously, from the point of view of light emission, heat energy represents so much wasted energy.

The ratio of energy radiated out in the form of light to total energy radiated out by the hot body is called the ***radiant*** efficiency of the luminous source and, obviously, depends on the temperature of the source. As the temperature is increased beyond that at which light waves were first given off, the radiant efficiency increases, because light energy will increase in greater proportion than the total radiated energy. When emitted light becomes white *i.e.*, it includes all the visible wavelengths, from extreme red to extreme violet, then a further increase in temperature produces radiations which are of wavelength smaller than that of violet radiations. Such radiations are invisible and are known as ultra-violet radiations. It is found that maximum radiant efficiency would occur at about 6200°C and even then the value of this maximum efficiency would be 20%. Since this temperature is far above the highest that has yet been obtained in practice, it is obvious that the actual efficiency of all artificial sources of light *i.e*. those depending on ***temperature incandescence***, is low.

As discussed above, light is radiant energy which is assumed to be propagated in the form of transverse waves through an invisible medium known as ether. These light waves travel with a velocity of 2.99776×10^8 m/s or 3×10^8 m/s approximately but their wavelengths are different. The wavelength of red light is nearly 0.000078 cm and that of violet light 0.000039 cm. Since these wavelengths are very small, instead of using 1 cm as the unit for their measurement, a submultiple 10^{-8} cm is used. This submultiple is known as Angstrom Unit (A.U.)

$$\mathbf{1\ A.U. = 10^{-8} cm = 10^{-10} m}$$

Hence, the wave-length of red light becomes $\lambda_r = 7800 \times 10^{-10}$ m or 7800 A.U. and $\lambda_v = 3900 \times 10^{-10}$ m or 3900 A.U. The sensation of colour is due to the difference in the wavelengths and hence frequencies of the light radiations.

Solid Angle:

Consider an area *A* which is part of a sphere of radius *r* (Fig. 4.1). Let us find the solid angle ω subtended by this area at the centre *C* of the sphere. For this purpose, let point *C* be joined to every point on the edges of the area *A*. Then, the angle enclosed by the cone at point *C* gives the solid angle.

Its value is $\boldsymbol{\omega = \frac{A}{r^2}}$ steradian

The unit of solid angle is ***steradian*** (sr). If, in the above equation, $A = r^2$, then ω = 1 steradian. Hence, steradian is defined as the angle subtended at the centre of a sphere by a part of its surface having an area equal to $(\text{radius})^2$.

Obviously, the solid angle subtended at the centre by whole of the spherical surface = $4\pi\, r^2/r^2 = 4\pi$ steradian (sr).

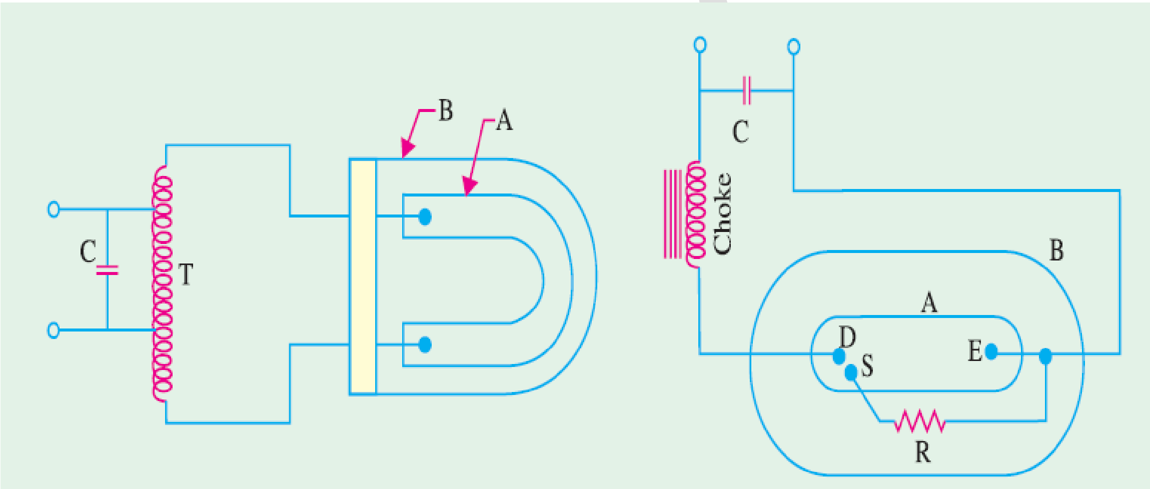


Fig. 4.1

Definitions:

Before proceeding further, definitions of a few principal terms employed in connection with illumination, are given below :

1. Candela. It is the unit of luminous intensity of a source. It is defined as 1/60th of the luminous intensity per cm^2 of a black body radiator at the temperature of solidification of platinum (2045°K).

A source of one candela (*cd*) emits one lumen per steradian. Hence, total flux emitted by it allround is $4\pi \times 1 = 4\pi$ lumen.

2. Luminous Flux (F or Φ). It is the light energy radiated out per second from the body in the form of luminous light waves. Since, it is a rate of flow of energy, it is a sort of *power* unit. Unit of luminous flux is *lumen* (lm). It is defined as the ***flux contained per unit solid angle of a source of one candela or standard candle*** (Fig. 4.2).

Approximate relation between lumen and electric unit of power
i.e. watt is given as

1 lumen = 0.0016 watt (approx.)

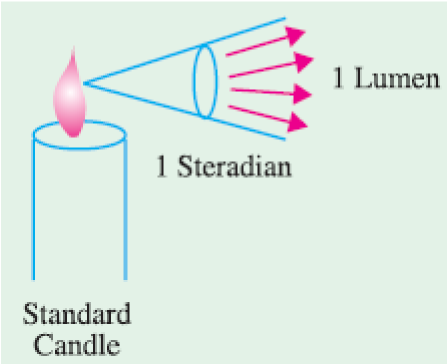


Fig 4.2

3. Lumen-hour. It is the quantity of light delivered in one hour by a flux of one lumen.

4. Luminous Intensity (I) or Candle-power of a point source in any particular direction is given by the ***luminous flux radiated out per unit solid angle in that direction***. In other words, it is solid angular flux density of a source in a specified direction.

If $d\Phi$ is the luminous flux radiated out by a source within a solid angle of $d\omega$ steradian in any particular direction, then $I = d\Phi/d\omega$.

If flux is measured in lumens and solid angle in steradian, then its unit is lumen/steradian (lm/sr) or candela (*cd*).

If a source has an average luminous intensity of I lm/sr (or I candela), then total flux radiated by it all around is $\Phi = \omega I = 4\pi I$ lumen.

Generally, the luminous intensity or candle power of a source is different in different directions. The average candle-power of a source is the average value of its candle power in all directions. Obviously, it is given by total flux (in lm) emitted in all directions in all planes divided by 4π. This average candle-power is also known as ***mean spherical candle-power*** (M.S.C.P.).

$$\therefore \mathbf{M.S.C.P.} = \frac{\text{total flux in lumens}}{4\pi}$$

If the average is taken over a hemisphere (instead of sphere), then this average candle power is known as ***mean hemispherical candle-power*** (M.H.S.C.P.).

It is given by the total flux emitted in a hemisphere (usually the lower one) divided by the solid angle subtended at the point source by the hemisphere.

$$\therefore \mathbf{M.H.S.C.P.} = \frac{\text{flux emitted in a hemisphere}}{2\pi}$$

5. Reduction Factor of a source is given by the ratio, f = M.S.C.P./M.H.C.P. where M.H.C.P. is the mean horizontal candle power.

It is also referred to as spherical reduction factor.

6. Illuminance or Illumination (E). When the luminous flux falls on a surface, it is said to be illuminated. The illumination of a surface is measured by the normal luminous flux per unit area received by it.

If $d\Phi$ is the luminous flux incident normally on an area dA, then $E = d\Phi/dA$ or $E = \Phi/A$.

Unit. Since flux Φ is measured in lumens and area in m^2, unit of E is lm/m^2 or lux. The alternative name is metre-candle (m-cd). Let us see why? Imagine a sphere of radius of one metre around a point source of one candela. Flux radiated out by this source is 4π lumen. This flux falls normally on the curved surface of the sphere which is $= 4\pi m^2$. Obviously, illumination at every point on the inner surface of this sphere is $4\pi\ lm/4\pi\ m^2 = 1\ lm/m^2$. However, the term lm/m^2 is to be perferred to metre-candle.

7. Luminance (L) of an Extended Source. Suppose ΔA is an element of area of an ***extended*** source and Δt its luminous intensity when viewed in a direction making an angle φ with the perpendicular to the surface of the source (Fig. 4.3), then luminance of the source element is given by

$$L = \frac{\Delta I}{\Delta A \cos\phi} = \frac{\Delta I}{\Delta A'}\ cd/m^2 \quad ...(i)$$

where $\Delta A' = \Delta A \cos\phi$

= area of the source element projected onto a plane perpendicular to the specified direction.

As will be seen from Art. 49.5.

$$E = \frac{I\cos\theta}{d^2} \quad \text{or} \quad \Delta E = \frac{\Delta I}{d^2}\cos\theta$$

Substituting the value of ΔI from Eq. (i) above, we get

$$\Delta E = \frac{L.\Delta A'}{d^2}\cos\theta = L\cos\theta.\,d\omega$$

where $d\omega = \Delta A'/d^2$ steradian

$$E = \int L\cos\theta\,.\,d\omega = L\int\cos\theta\,.\,d\omega$$

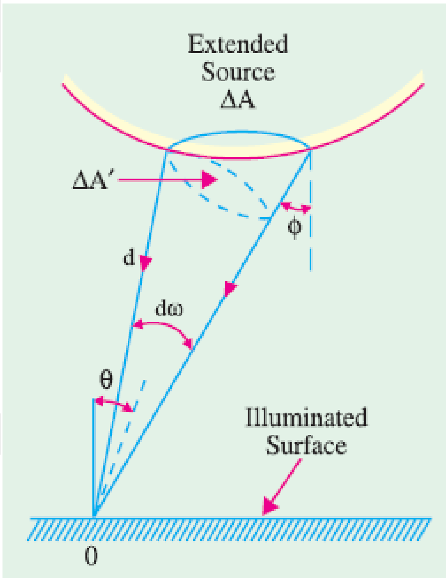


Fig 4.3

8. Luminous Exitance (M) of a Surface. The luminous exitance (M) at a point on a surface is defined as luminous flux emitted per unit area in all directions. If an element of an illuminated area ΔA emits a total flux of $\Delta\Phi$ in all directions (over a solid angle of 2π steradian) then

$$M = \Delta\Phi/\Delta A \quad lm/m^2$$

It can be proved that $M = \pi L$ in the case of a uniform diffuse ***source***.

9. Transmittance (T) of an Illuminated Diffuse Reflecting Surface. It is defined as the ratio of the total luminous flux transmitted by it to the total flux incident on it.

The relation between luminous exitance (M) of a surface transmitting light and illuminance (E) on the other side of it is

$$\boldsymbol{M = T\,E} \text{ or } \boldsymbol{T = M/E}$$

Since light falling on a surface is either transmitted, reflected or absorbed the following relation holds good

$\boldsymbol{T + \rho + \alpha = 1}$ where α is the absorptance of the surface.

10. Reflection Ratio or Coefficient of Reflection or Reflectance (ρ). It is given by the luminous flux reflected from a small area of the surface to the total flux incident upon it

$\rho = M/E$ *i.e.* ratio of luminous exitance and illuminance.

It is always less than unity. Its value is zero for ideal 'black body' and unity for a perfect reflector.

11. Specific Output or Efficiency of a lamp is the ratio of luminous flux to the power intake. Its unit is lumen/watt (lm/W). Following relations should be taken note of :

(*a*) $$\frac{\text{lumen}}{\text{watt}} = \frac{4\pi \times \text{M.S.C.P.}}{\text{watt}}$$

or $$\frac{\text{lm}}{\text{W}} = \frac{4\pi}{\text{watt/M.S.C.P.}}$$

(*b*) since $f = \text{M.S.C.P./M.H.C.P.}$ $\therefore$ $\text{lm/W} = \dfrac{4\pi f}{\text{watt/M.S.C.P.}}$

(*c*) Obviously, $$\text{watts/M.S.C.P.} = \frac{4\pi}{\text{lm/W}} = \frac{\text{watt/M.H.C.P.}}{f}$$

(*d*) Also $$\text{watts/M.H.C.P.} = \frac{4\pi f}{\text{lm/W}} = f \times \text{watts/M.S.C.P.}$$

12. Specific Consumption. It is defined as the ratio of the power input to the average candlepower. It is expressed in terms of watts per average candle or watts/M.S.C.P.

The summary of the above quantities along with their units and symbol is given in Table 4.1.

Table 4.1.

Name of Qty	Unit	Symbols
Luminous Flux	Lumen	F or Φ
Luminous Intensity (candle-power)	Candela	I
Illumination or Illuminance	lm/m^2 or lux	E
Luminance or Brightness	cd/m^2	L or B
Luminous Exitance	lm/m^2	M

Laws of Illumination or Illuminance:

The illumination **(*E*)** of a surface depends upon the following factors. The source is assumed to be a point source or is otherwise sufficiently away from the surface to be regarded as such.

(i) ***E*** is directly proportional to the luminous intensity **(*I*)** of the source or $E \propto I$

(ii) ***Inverse Square Law***. The illumination of a surface is inversely proportional to the square of the distance of the surface from the source.

In other words $E \propto 1/r^2$

Proof

In Fig. 4.4 are shown portions of the surfaces of three spheres whose radii are in the ratio 1 : 2 : 3. All these portions, obviously, subtend the same solid angle at the source and hence receive the same amount of total flux. However, since their areas are in the ratio of 1 : 4 : 9, their illuminations are in the ratio $1 : \frac{1}{4} : \frac{1}{9}$

(*iii*) *Lambert's Cosine Law*. According to this law, *E* is directly proportional to the cosine of the ***angle made by the normal to the illuminated surface with the direction of the incident flux***.

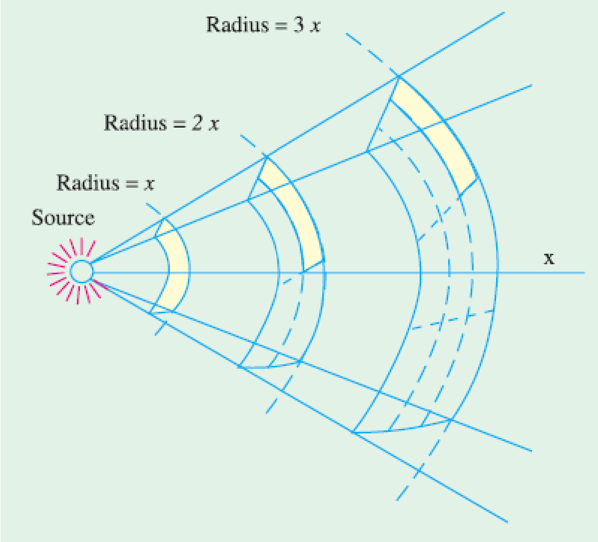


Fig. 4.4

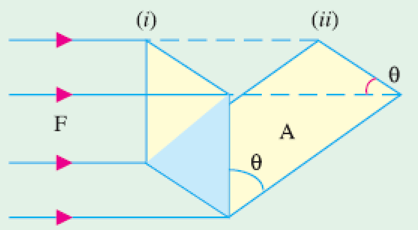


Fig.4.5

Proof

As shown in Fig. 4.5, let Φ be the flux incident on the surface of area *A* when in position 1. When this surface is turned back through an angle θ, then the flux incident on it is Φ cos θ. Hence, illumination of the surface when in position 1 is $E_1 = \Phi/A$. But when in position 2.

$$E_2 = \frac{\Phi \cos\theta}{A} \qquad \therefore \qquad E_2 = E_1 \cos\theta$$

Combining all these factors together, we get $E = I \cos\theta/r^2$. The unit is lm/m^2.

The above expression makes the determination of illumination possible at a given point provided the position and the luminous intensity or candle power (in the given direction) of the source (or sources) by which it is illuminated are known as illustrated by the following examples.

Consider a lamp of uniform luminous intensity suspended at a height *h* above the working plane as shown in Fig. 4.6. Let us consider the value of illumination at point *A* immediately below the lamp and at other points *B*,*C*,*D* etc., lying in the working plane at different distances from *A*.

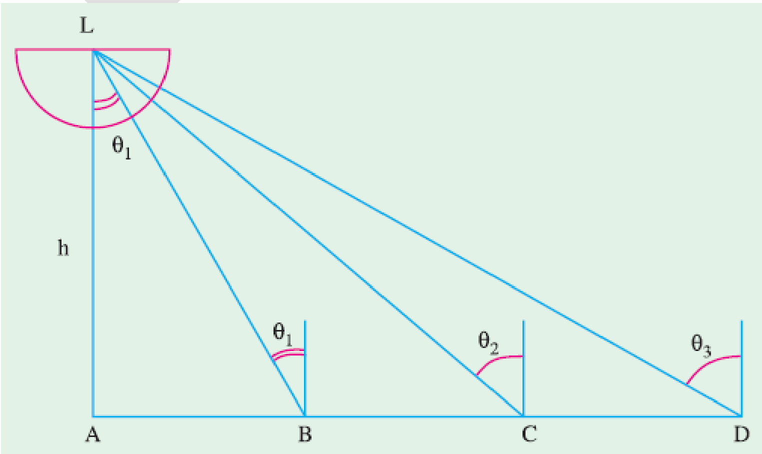


Fig.4.6

$$E_A = \frac{I}{h^2} \text{ —since } \theta = 0 \text{ and } \cos\theta = 1$$

$$E_B = \frac{I}{LB^2} \times \cos\theta_1. \quad \text{Since,} \quad \cos\theta_1 = h/LB$$

$$\therefore \quad E_B = \frac{I}{LB^2} \times \frac{h}{LB} = I \times \frac{h}{LB^2} = \frac{1}{h^2} \cdot \frac{h^3}{LB^3} = \frac{1}{h^2}\left(\frac{h}{LB}\right)^3$$

$$\text{Now} \quad \frac{1}{h^2} = E_A \text{ and } \left(\frac{h}{LB}\right)^3 = \cos^3\theta_1$$

$$\therefore \quad E_B = E_A \cos^3\theta_1$$

Similarly, $E_C = E_A . \cos^3\theta_2$ and $E_D = E_A \cos^3\theta_3$ and so on.

Example. *A lamp giving out 1200 lm in all directions is suspended 8 m above the working plane. Calculate the illumination at a point on the working plane 6 m away from the foot of the lamp.*

Solution. Luminous intensity of the lamp is

$I = 1200/4\pi = 95.5$ cd

As seen from Fig. 4.7.

$$L_B = \sqrt{8^2 + 6^2} = 10 \text{ m} ; \cos\theta = 8/10 = 0.8$$

Now, $E = I \cos\theta / r^2$

$\therefore E_B = 95.5 \times 0.8/10^2 =$ **0.764 lm/m²**

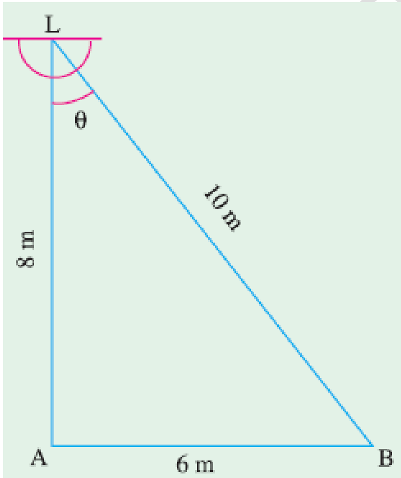


Fig.4.7

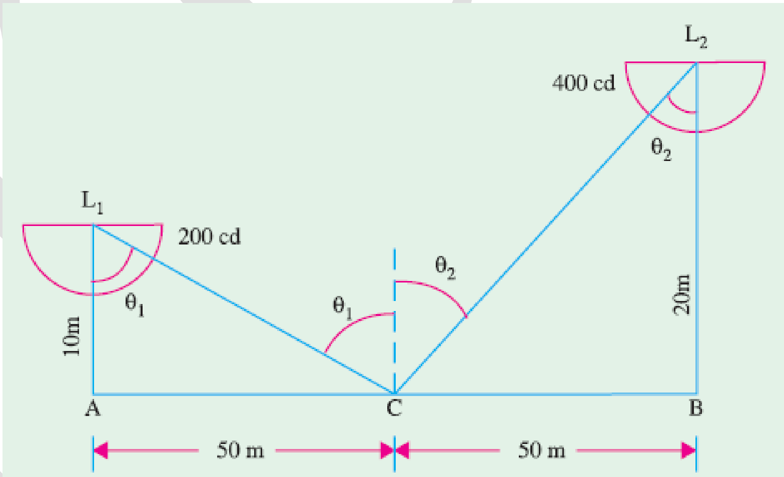


Fig.4.8

Example. *Two lamps A and B of 200 candela and 400 candela respectively are situated 100 m apart. The height of A above the ground level is 10 m and that of B is 20 m. If a photometer is placed at the centre of the line joining the two lamp posts, calculate its reading.*

Solution. When the illumination photometer is placed at the centre point, it will read the value of combined illumination produced by the two lamps (Fig. 4.8).

$$\text{Now,}\quad L_1C = \sqrt{10^2 + 50^2}$$
$$= 51 \text{ m}$$
$$L_2C = \sqrt{20^2 + 50^2}$$
$$= 53.9 \text{ m}$$
$$\cos\theta_1 = 10/51\ ;$$
$$\cos\theta_2 = 20/53.9$$

Illumination at point C due to lamp L_1

$$= \frac{200 \times 10}{51 \times 2600}$$
$$= \mathbf{0.015\ lm/m^2}$$

Similarly, illumination due to lamp L_2

$$= \frac{400 \times 20/53.9}{2900} = \mathbf{0.051\ lm/m^2}$$

$$\therefore \quad E_C = 0.015 + 0.051$$
$$= \mathbf{0.066\ lm/m^2\ or\ lux}$$

Laws Governing Illumination of Different Sources:

The laws applicable to the illumination produced by the following three types of sources will be considered.

(*i*) Point Source

As discussed above, the law governing changes in illumination due to point source of light is $E = I \cos\theta/d^2$.

(*ii*) Line Source

Provided the line source is of infinite length and of uniform intensity, the illumination at a point lying on a surface parallel to and facing the line source is given by

$$E = \frac{\pi I}{2d} \text{ lm/m}^2$$

where I = luminous intensity normal to the line source (in candles per-meter length of the sources).

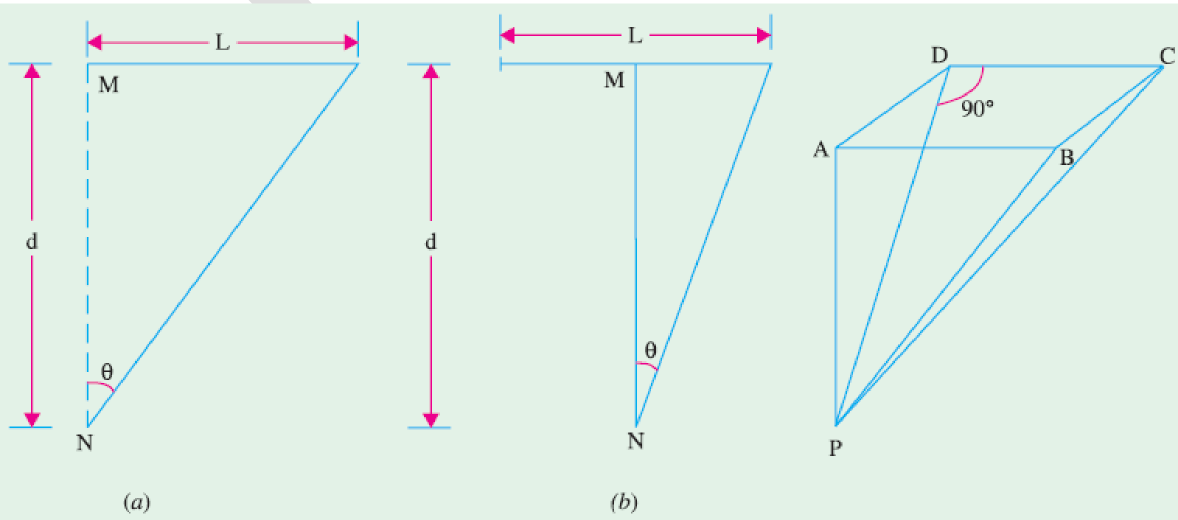


Fig. 4.9 Fig. 4.10

However, in practice, the line sources are of finite length, so that the following law applies.

$$E = \frac{I}{4d}(\sin 2\theta + 2\theta)\ \text{lm/m}^2$$
$$= \frac{I}{2d}(\sin 2\theta + 2\theta)\ \text{lm/m}^2$$

—Fig. 4.9 (*a*) and Fig. 4.9 (b).

where I = candle power per metre length in a direction normal to the line source.

$$= \frac{\Phi}{\pi^2 L} \quad \text{cd/m}$$

where Φ is the total flux of the source in lumens and L is the length of the line source in metres.

Polar Curves of C.P. Distribution:

All our calculations so far were based on the tacit assumption that the light source was of equal luminous intensity or candle-power in all directions. However, lamps and other sources of light, as a rule, do not give uniform distribution in the space surrounding them. If the actual luminous intensity of a source in various directions be plotted to scale along lines radiating from the centre of the source at corresponding angles, we obtain the polar curve of the candle power.

Suppose we construct a figure consisting of large number of spokes radiating out from a point, the length of each spoke representing to some scale the candle power or luminous intensity of the source in that particular direction. If now we join the ends of these spokes by some suitable material, say, by linen cloth, then we get a surface whose shape will represent to scale the three dimensional candle power distribution of the source placed at the centre.
In the ideal case of a point source having equal distribution in all directions, the surface would be spherical.

It would be realized that it is difficult to give a graphic representation of such a 3-dimensional model in a plane surface. Therefore, as with engineering drawings, it is usual to draw only one or more elevations and a plan of sections through the centre of the source. Elevations represent c.p. distribution in the ***vertical*** plane and the plans represent c.p. distribution in ***horizontal*** plane. The number of elevations required to give a complete idea of the c.p. distribution of the source in all directions depends upon the shape of the plan *i.e.* on the horizontal distribution. If the distribution is uniform in every horizontal plane *i.e.* if the polar curve of horizontal distribution is a circle, then only one vertical curve is sufficient to give full idea of the space distribution.

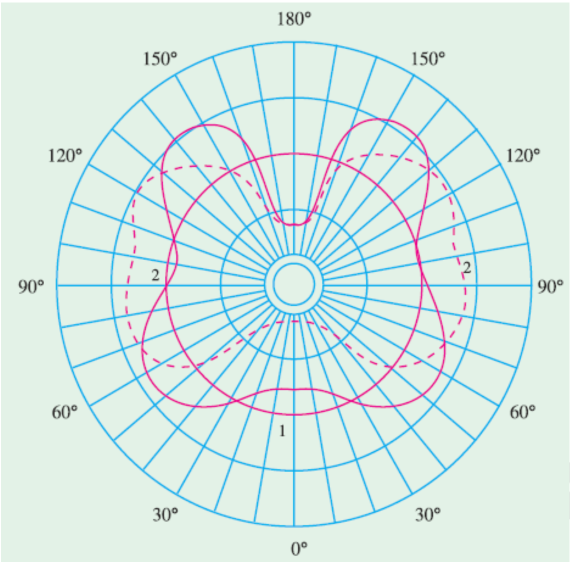


Fig. 4.11

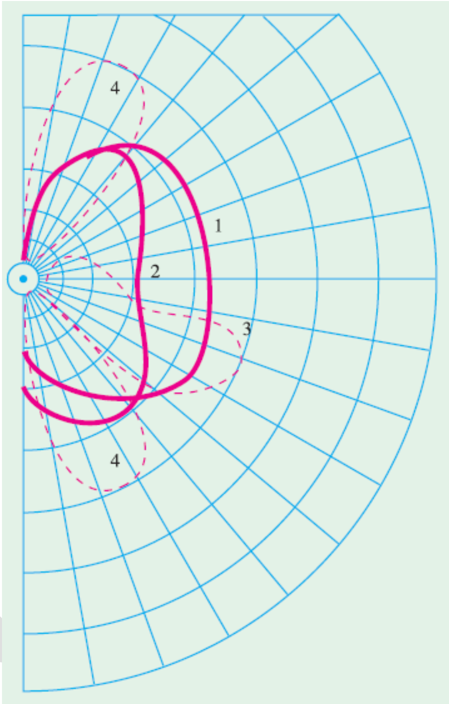


Fig. 4.12

In figure 4.11 are shown two polar curves of c.p. distribution in a vertical plane. Curve 1 is for vacuum type tungsten lamp with zig-zag filament whereas curve 2 is for gas filled tungsten lamp with filament arranged as a horizontal ring.

If the polar curve is symmetrical about the vertical axis as in the figures given below, then it is sufficient to give only the polar curve within one semicircle in order to completely define the distribution of c.p. as shown in figure 4.12.

The curves 1 and 2 are as in figure 4.11, curve 3 is for d.c. open arc with plain carbons and curve 4 is for a.c. arc with plain carbons. However, if the source and/or reflector are not symmetrical about vertical axis, it is impossible to represent the space distribution of c.p. by a single polar diagram and even polar diagrams for two planes at right angles to each other give no definite idea as to the distribution in the intermediate planes.

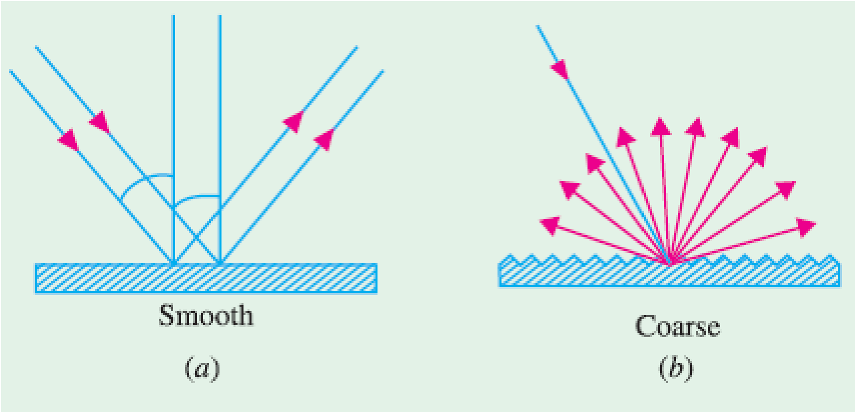
Uses of Polar Curves:

Polar curves are made use of in determining the M.S.C.P. etc. of a source. They are also used in determining the actual illumination of a surface *i.e.* while calculating the illumination in a particular direction, the c.p. in that particular direction as read from the vertical polar curve, should be employed.

Diffusing and Reflecting Surfaces : Globes and Reflectors

When light falls on polished metallic surfaces or silvered surfaces, then most of it is reflected back according to the laws of reflection i.e. the angle of incidence is equal to the angle of reflection. Only a small portion of the incident light is absorbed and there is always the image of the source. Such reflection is known as ***specular*** reflection.

However, as shown in figure, if light is incident on coarse surfaces like paper, frosted glass, painted ceiling etc., then it is scattered or diffused in all directions, hence no image of the source is formed. Such reflection of light is called ***diffuse reflection***. A perfect diffuser is one that scatters light uniformly in all directions and hence appears equally bright from whatever direction it is viewed. A white blotting paper is the nearest approach to a diffuser.



By reflecting factor of a surface is meant the ratio

$$= \frac{\text{reflected light}}{\text{incident light}}$$

It is also known as reflection ratio or coefficient of reflection of a surface.

If the light is incident on a transparent surface, then some of it is absorbed and greater percentage of it passes through and emerges on the other side.

To avoid direct glare from electric arcs and incandescent filament lamps, they are surrounded more or less completely by diffusing shades or globes. In addition, a reflector may also be embodied to prevent the escape of light in directions where it serves no useful purpose. In that case, so far as the surroundings are concerned, the diffusing globe is the source of light. Its average brilliancy is lower the more its diffusing area. Depending on the optical density, these globes absorb 15 to 40% of light emitted by the encircled bulb. The bulbs may also be frosted externally by etching or sand-blasting but internal frosting is better because there is no sharp scratching or cracks to weaken the glass.

In domestic fittings, a variety of shades are used whose main purpose is to avoid glare. Properly designed and installed prismatic glass shades and holophane type reflectors have high efficiency and are capable of giving accurate predetermined distribution of light.

Regular metallic reflection is used in search-light mirrors and for general lighting purposes. But where it is used for general lighting, the silvered reflectors are usually fluted to make the illumination as uniform as possible.

Regular cleaning of all shades, globes, and reflectors is very important otherwise the loss of light by absorption by dust etc., collected on them becomes very serious.

Various types of reflectors are illustrated in figures (*i*) to (*v*). Figure (*i*) shows a holophane stiletto reflector used where extensive, intensive or focussing light distribution is required.

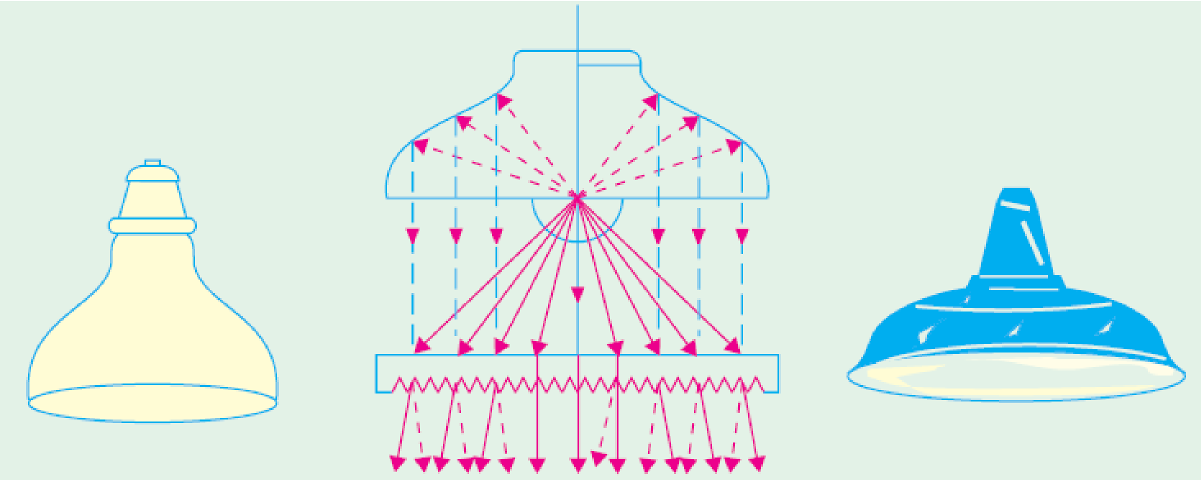


Fig. (*i*) Fig. (*ii*) Fig. (*iii*)

The optical combination of a lamp, reflector and a lens plate, as shown in Fig. (*ii*), provides a high degree of light control. Multiple panels can be conveniently incorporated in fittings suited to different architectural schemes. The dispersive reflector of Fig. (*iii*) is suitable for practically all classes of industrial installations. The reflector is a combination of concave and cylindrical reflecting surfaces in the form of a deep bowl of wide dispersive power. It gives maximum intensity between 0º and 45º from the vertical.

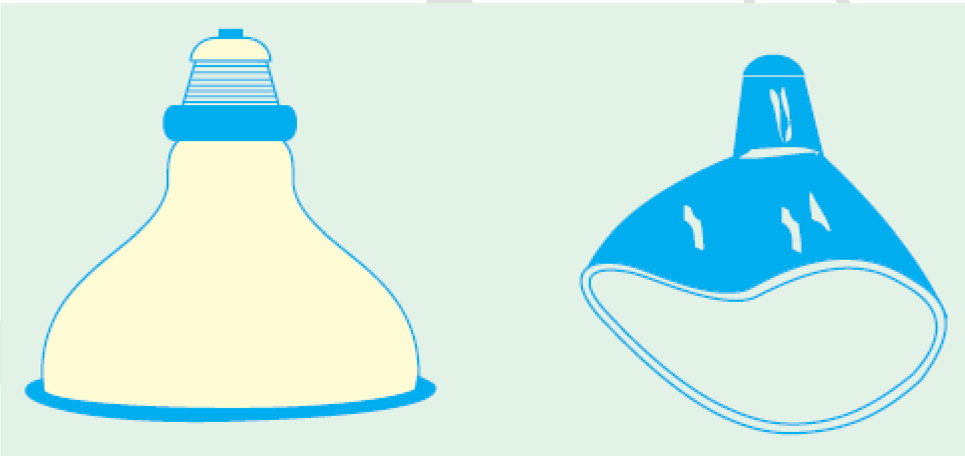


Fig. (*iv*) Fig. (*v*)

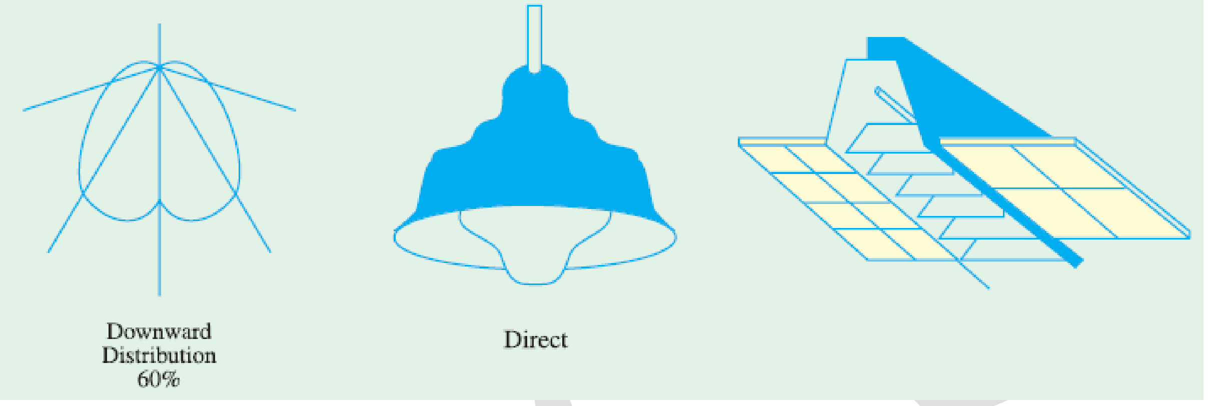
The concentrating reflector of parabolic form shown in Fig. (*iv*) is suitable for situations requiring lofty installations and strongly-concentrated illumination as in public halls, foundries and power stations etc. They give maximum intensity in zones from 0º to 25º from the vertical. The elliptical angle reflector shown in Fig. (*v*) is suitable for the side lighting of switchboards, show windows etc., because they give a forward projection of light in the vertical plane and spread the light in the horizontal plane.

Lighting Schemes:

Different lighting schemes may be classified as ***(i)*** direct lighting ***(ii)*** indirect lighting and ***(iii)*** semi-direct lighting ***(iv)*** semi-indirect lighting and ***(v)*** general diffusing systems.

(*i*) Direct Lighting:

As the name indicates, in the form of lighting, the light from the source falls directly on the object or the surface to be illuminated (Fig. below). With the help of shades and globes and reflectors of various types, most of the light is directed in the lower hemisphere and also the brilliant source of light is kept out of the direct line of vision. Direct illumination by lamps in suitable reflectors can be supplemented by standard or bracket lamps on desk or by additional pendant fittings over counters.

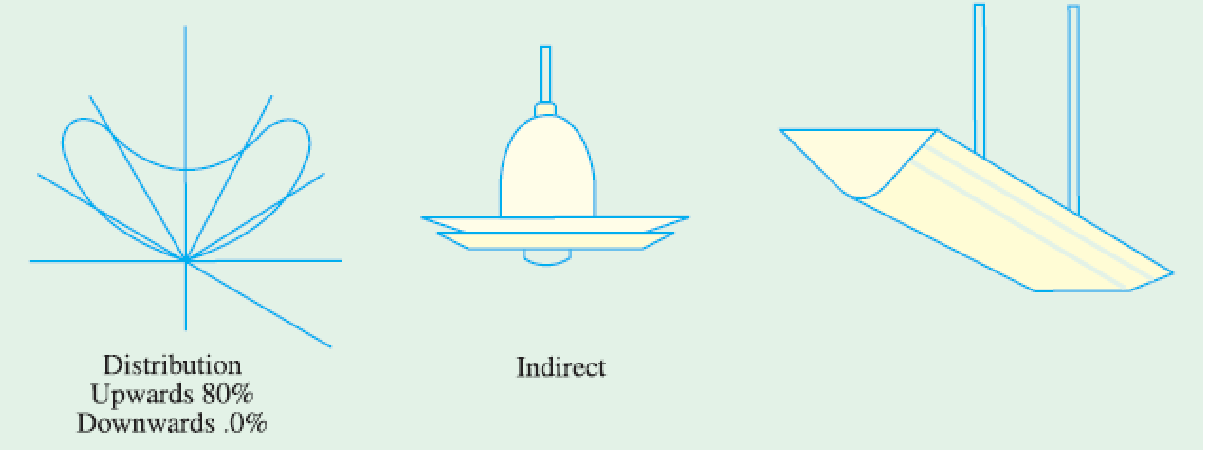


The fundamental point worth remembering is planning any lighting installation is that sufficient and sufficiently uniform lighting is to be provided at the working or reading plane. For this purpose, lamps of suitable size have to be so located and furnished with such fittings as to give correct degree and distribution of illumination at the required place. Moreover, it is important to keep the lamps and fittings clean otherwise the decrease in effective illumination due to dirty bulbs or reflectors may amount to 15 to 25% in offices and domestic lighting and more in industrial areas as a result of a few weeks neglect.

Direct lighting, though most efficient, is liable to cause glare and hard shadows.

(*ii*) Indirect Lighting:

In this form of lighting, light does not reach the surface directly from the source but indirectly by diffuse reflection (Fig. below). The lamps are either placed behind a cornice or in suspended ***opaque*** bowls. In both cases, a silvered reflector which is corrugated for eliminating striations is placed beneath the lamp.



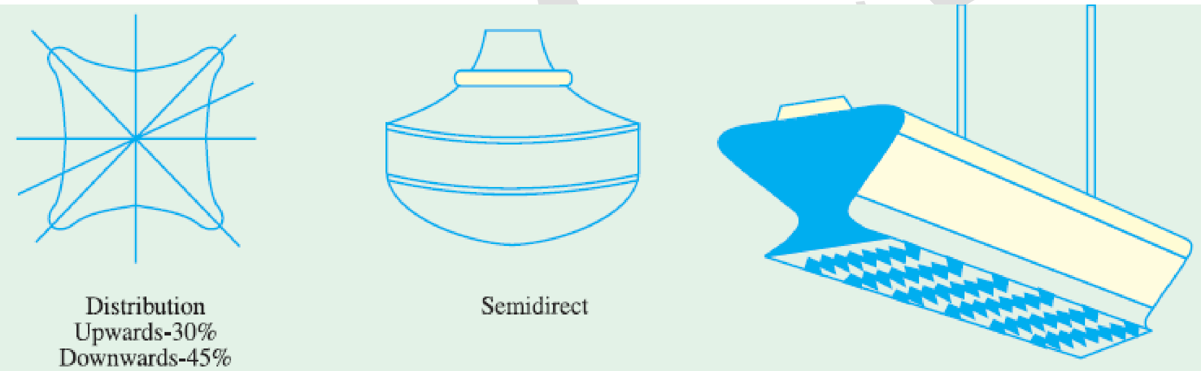
In this way, maximum light is thrown upwards on the ceiling from which it is distributed all over the room by diffuse reflection. Even gradation of light on the ceiling is secured by careful adjustment of the position and the number of lamps. In the cornice and bowl system of lighting, bowl fittings are generally suspended about three-fourths the height of the room and in the case of cornice lighting, a frieze of curved profile aids in throwing the light out into the room to be illuminated. Since in indirect lighting whole of the light on the working plane is received by diffuse reflection, it is important to keep the fittings clean.

One of the main characteristics of indirect lighting is that it provides shadowless illumination which is very useful for drawing offices, composing rooms and in workshops especially where large machines and other obstructions would cast troublesome shadows if direct lighting were used.

However, many people find purely indirect lighting flat and monotonous and even depressive. Most of the users demand 50 to 100% more light at their working plane by indirect lighting than with direct lighting. However, for appreciating relief, a certain proportion of direct lighting is essential.

(*iii*) Semi-direct System

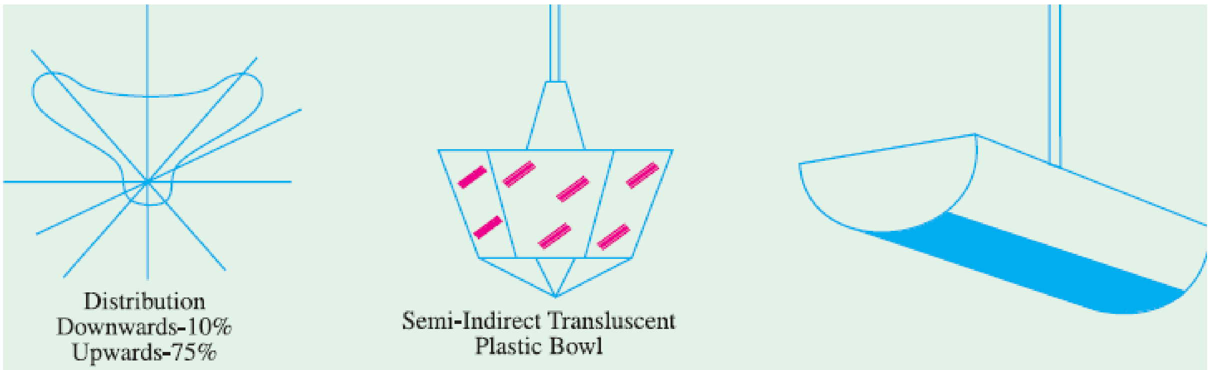
This system utilizes luminaries which send most of the light downwards directly on the working plane but a considerable amount reaches the ceilings and walls also (Fig. below).



The division I s usually 30% upwards and 45% downwards. Such a system is best suited to rooms with high ceilings where a high level of uniformly-distributed illumination is desirable. Glare in such units is avoided by using diffusing globes which not only improve the brightness towards the eye level but improve the efficiency of the system with reference to the working plane.

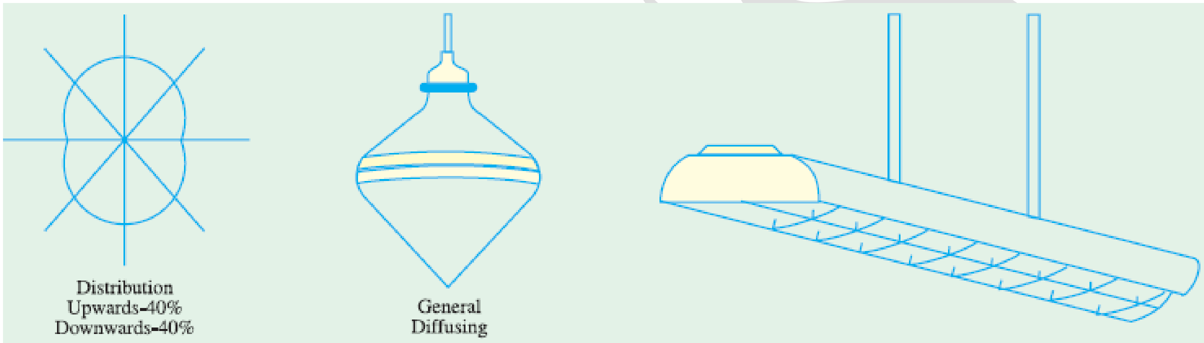
(*iv*) Semi-indirect Lighting

In this system which is, in fact, a compromise between the first two systems, the light is partly received by diffuse reflection and partly direct from the source (Fig. below). Such a system, therefore, eliminates the objections of indirect lighting mentioned above. Instead of using opaque bowls with reflectors, translucent bowls without reflector are used. Most of the light is, as before, directed upwards to the ceiling for diffuse reflection and the rest reaches the working plane directly except for some absorption by the bowl.



(*v*) General Diffusing System

In this system, luminaries are employed which have almost equal light distribution downwards and upwards as shown in Fig. below.



Illumination Required for Different Purposes:

There has been a steady movement towards higher intensities for artificial illumination during the last few decades. The movement is likely to continue because the highest intensities in average installations are much less than those of the diffused daylight. The human eye possesses a tremendous power of accommodation and it can work comfortably within an enormous range of illuminations.

For example, at full noon, sun provides about 120,000 lm/m^2, diffuse day-light near a window is of the order of 600 lm/m^2 (value varying widely) and full moon-light gives 0.1 to 0.3 lm/m^2. For reading, usually 20 to 30 lm/m^2 is generally considered sufficient, though daylight illumination is much higher.

Some persons can read without much strain even when illumination is as low as 3 lm/m^2. Because of this, it is difficult to lay down definite values of illumination for various purposes but the following summary will be found useful :

Purpose and Places	***lm/m^2***
Precision work, displays, tasks requiring rapid discrimination	above 500
Extra fine machine work, around needles of sewing machines, fine engraving, inspection of fine details having low contrast	200-500
Proof-reading, drawing, sustained reading, fine assembling, skilled bench-work	100-200

Drawing offices, art exhibition, usual reading	60-100
In museums, drill halls, for work of simple nature not involving close attention to fine details	40-60
Usual observation as in bed-rooms, waiting rooms, auditoriums and general lighting in factories	20-40
Hospital wards, yards, railway platforms and corridors	5-10

Space/Height Ratio:

It is given by the ratio : $\dfrac{\text{horizontal distance between two lamps}}{\text{mounting height of lamps}}$

This ratio depends on the nature of the polar curve of a lamp ***when used along with its reflector***. A reflector has tremendous influence on the shape of the polar curve of the lamp, hence the value of space/height ratio, in fact, depends entirely on the type of reflector used. For obtaining uniform illumination on the working plane, it is essential to choose a correct value for this ratio.

In other words, it means that a reflector gives uniform illumination for a definite value of this ratio only. The ratio may be found easily if the polar curve of the type of fixture used is known. For reflectors normally used in indoor lighting, the value of this ratio lies between 1 and 2.

Design of Lighting Schemes and Lay-outs:

A well-designed lighting scheme is one which

(*i*) provides adequate illumination

(*ii*) avoids glare and hard shadows

(*iii*) provides sufficiently uniform distribution of light all over the working plane.

Before explaining the method of determining the number, size and proper arrangement of lamps in order to produce a given uniform illumination over a certain area, let us first consider the following two factors which are of importance in such calculations.

Utilization Factor or Coefficient of Utilization (η):

It is the ratio of the lumens actually received by a particular surface to the total lumens emitted by a luminous source.

$$\therefore\ \eta = \frac{\text{lumens actually recieved on working plane}}{\text{lumens emitted by the light source}}$$

The value of this factor varies widely and depends on the following factors :

1. the type of lighting system, whether direct or indirect etc.

2. the type and mounting height of the fittings

3. the colour and surface of walls and ceilings and

4. to some extent on the shape and dimensions of the room.

For example, for direct lighting, the value of η varies between 0.4 and 0.6 and mainly depends on the shape of the room and the type and mounting height of fittings but very little on the colour of walls and ceiling. For indirect lighting, its value lies between 0.1 and 0.35 and the effect of walls and ceiling, from which light is reflected on the working plane, is much greater. Exact determination of the value of utilization factor is complicated especially in small rooms where light undergoes multiple reflections.

Since the light leaving the lamp in different directions is subjected to different degrees of absorption, the initial polar curve of distribution has also to be taken into account. Even though manufacturers of lighting fittings supply tables giving utilization factors for each type of fitting under specified conditions yet, since such tables apply only to the fittings for which they have been compiled, a good deal of judgment is necessary while using them.

Depreciation Factor (p):

This factor allows for the fact that effective candle power of all lamps or luminous sources deteriorates owing to blackening and/or accumulation of dust or dirt on the globes and reflectors etc. Similarly, walls and ceilings etc., also do not reflect as much light as when they are clean. The value of this factor may be taken as 1/1.3 if the lamp fittings are likely to be cleaned regularly or 1/1.5 if there is much dust etc.

$$p = \frac{\text{illumunation under actual conditions}}{\text{illumination when everything is perfectly clean}}$$

Since illumination is specified in lm/m^2, the area in square metre multiplied by the illumination required in lm/m^2 gives the total useful luminous flux that must reach the working plane. Taking into consideration the utilization and depreciation or maintenance factors, the expression for the gross lumens required is

Total lumens,

$$\phi = \frac{E \times A}{\eta \times p}$$

where E = desired illumination in lm/m^2 ; A = area of working plane to be illuminated in m^2
p = depreciation or maintenance factor ; η = utilization factor.

The size of the lamp depends on the number of fittings which, if uniform distribution is required, should not be far apart. The actual spacing and arrangement is governed by space/height values and by the layout of ceiling beams or columns. Greater the height, wider the spacing that may be used, although the larger will be the unit required. Having settled the number of units required, the lumens per unit may be found from (total lumens/number of units) from which the size of lamp can be calculated.

Example. *A room 8 m × 12 m is lighted by 15 lamps to a fairly uniform illumination of 100 lm/m^2. Calculate the utilization coefficient of the room given that the output of each lamp is 1600 lumens.*

Solution. Lumens emitted by the lamps = 15 × 1600 = 24,000 lm
Lumens received by the working plane of the room = 8 × 12 × 100 = 9600 lm
Utilization coefficient = 9600/24,000 = **0.4 or 40%.**

Example. *The illumination in a drawing office 30 m × 10 m is to have a value of 250 lux and is to be provided by a number of 300-W filament lamps. If the coefficient of utilization is 0.4 and the depreciation factor 0.9, determine the number of lamps required. The luminous efficiency of each lamp is 14 lm/W.*

Solution. $\Phi = EA/\eta p$; $E = 250$ lm/m^2, $A = 30 \times 10 = 300$ m^2; $\eta = 0.4$, $p = 0.9$

$\therefore \Phi = 250 \times 300/0.4 \times 0.9 = 208{,}333$ lm

Flux emitted/lamp = 300 × 14 = 4200 lm; No. of lamps reqd.= 208,333/4200 = **50.**

Example. *A drawing hall in an engineering college is to be provided with a lighting installation. The hall is 30 m × 20 m × 8 m (high). The mounting height is 5 m and the required level of illumination is 144 lm/m^2. Using metal filament lamps, estimate the size and number of single lamp luminaries and also draw their spacing layout. Assume : Utilization coefficient = 0.6; maintenance factor = 0.75; space/height ratio=1 lumens/watt for 300-W lamp = 13,lumens/watt for 500-W lamp = 16.*

Solution. Flux is given by $\Phi = EA/\eta\, p = 30 \times 20 \times 144/0.6 \times 0.75 = 192{,}000$ lm

Lumen output per 500-W lamp = 500 × 16 = 8,000

∴ No. of 500-W lamps required = 192,000/8000 = **24**

Similarly, No. of 300-W lamps required = 192,000/3900 = **49**

The 300-W lamps cannot be used because their number cannot be arranged in a hall of 30 m × 20 m with a space/height ratio of unity. However, 500-W lamps can be arranged in 4 rows of 6 lamps each with a spacing of 5 m both in the width and the length of the hall as shown in figure 4.13.

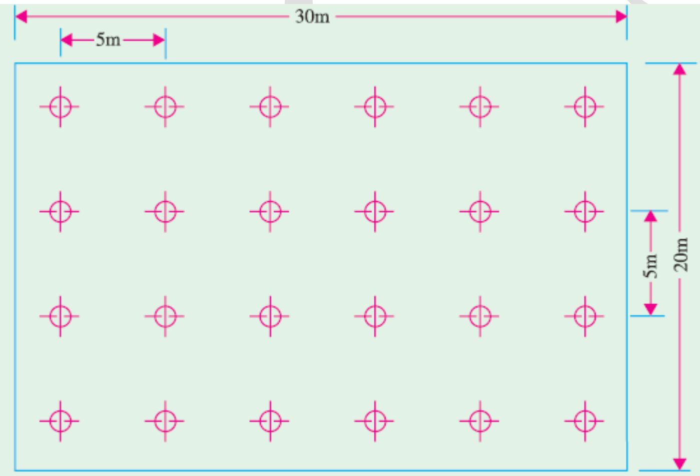


Fig. 4.13

Example. *Estimate the number and wattage of lamps which would be required to illuminate a workshop space 60 × 15 metres by means of lamps mounted 5 metres above the working plane. The average illumination required is about 100 lux. Coefficient of utilization=0.4 ; Luminous efficiency=16 lm/W. Assume a spacing/height ratio of unity and a candle power depreciation of 20%.*

Solution. Luminous flux is given by

$$\Phi = EA/\eta\, p = [100\ (60 \times 15)]/[\ 0.4 \times 1/1.2] = 27 \times 10^4 \text{ lm}$$

Total wattage reqd. = $27 \times 10^4/16$ = **17,000 W**

For a space/height ratio of unity, only three lamps can be mounted along the width of the room. Similarly, 12 lamps can be arranged along the length of the room.

Total number of lamps required is 12 × 3 = **36**.

Wattage of each lamp is = 17,000/36 = 472 W.

We will take the nearest standard lamp of **500 W**. These thirty-six lamps will be arranged as shown in figure 4.14.

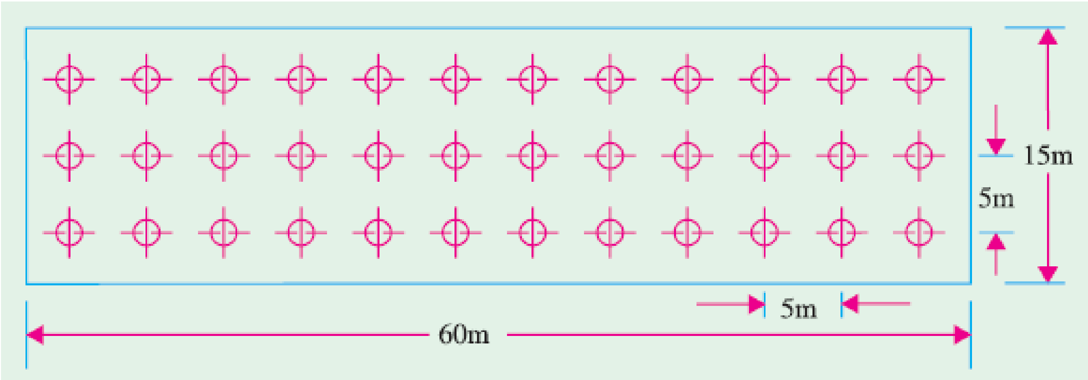


Fig. 4.14

Floodlighting:

It means 'flooding' of large surfaces with the help of light from powerful projectors. Flooding is employed for the following purposes :

1. For aesthetic purposes as for enhancing the beauty of a building by night *i.e.* floods lighting of ancient monuments, religious buildings on important festive occasions etc.
2. For advertising purposes *i.e.* flood lighting, huge hoardings and commercial buildings.
3. For industrial and commercial purposes as in the case of railway yards, sports stadiums and quarries etc.

Usually, floodlight projectors having suitable reflectors fitted with standard 250-, 500-, or 1,000-watt gas-filled tungsten lamps, are employed. One of the two typical floodlight installations often used is as shown in figure 4.15 (*a*). The projector is kept 15 m to 30 m away from the surface to be floodlighted and provides approximately parallel beam having beam spread of 25° to 30°. figure 4.15 (*b*) shows the case when the projector cannot be located away from the building. In that case, an asymmetric reflector is used which directs more intense light towards the top of the building.

The total luminous flux required to floodlight a building can be found from the relation, $\Phi = EA/\eta \times p$.

However, in the case of flood-lighting, one more factor has to be taken into account. That factor is known as waste-light factor (*W*). It is so because when several projectors are used, there is bound to be a certain amount of overlap and also because some light would fall beyond the edges of the area to be illuminated. These two factors are taken into account by multiplying the theoretical value of the flux required by a waste-light factor which has a value of nearly 1.2 for regular surfaces and about 1.5 for irregular objects like statues etc. Hence, the formula for calculation of total flux required for floodlighting purposes is

$$\phi = \frac{E \times A \times W}{\eta \times p}$$

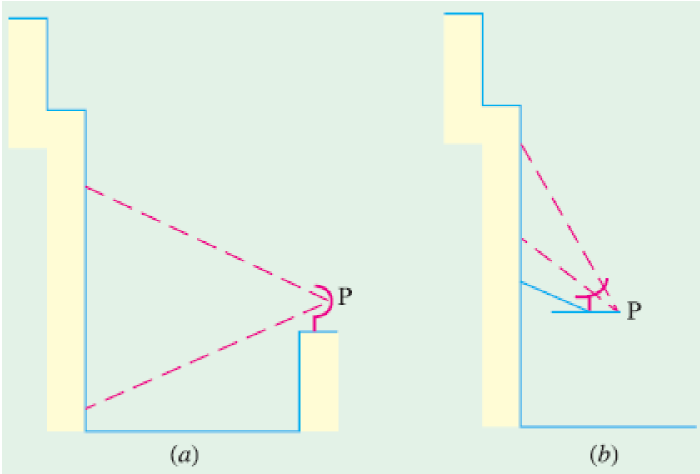


Fig. 4.15

Example. *It is desired to floodlight the front of a building 42 m wide and 16 m high. Projectors of 30° beam spread and 1000-W lamps giving 20 lumen/watt are available. If the desired level of illumination is 75 lm/m^2 and if the projectors are to be located at ground level 17 m away, design and show a suitable scheme. Assume the following :*
Coefficient of utilization = 0.4 ; Depreciation factor = 1.3; Waste-light factor = 1.2.
Solution.

$$\phi = \frac{E \times A \times W}{\eta \times p}$$

Here E = 75 lm/m^2 ; A = 42 × 16 = 672 m^2 ; W = 1.2; η = 0.4 ; p = 1/1.3

$$\phi = \frac{75 \times 672 \times 1.2}{0.4 \times 1/1.3} = 1{,}96{,}500\ lm$$

Lumen output of each 1,000-W lamp = 1,000 × 20 = 20,000 lm
No. of lamps required = 196,500/20,000 = **10.**
With a beam spread of 30°, it is possible to cover the whole length and width of the building by arranging the 10 projectors in two rows as shown in figure 4.16 (*a*).

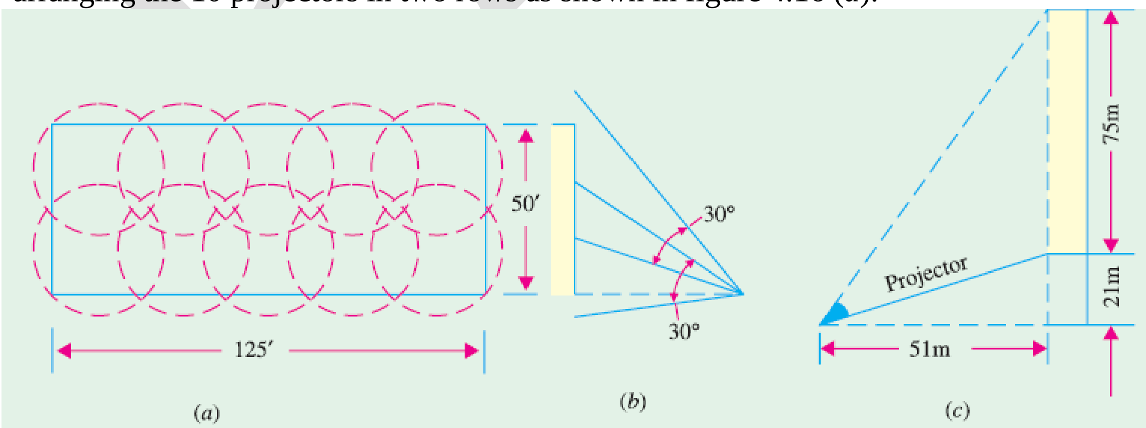


Fig. 4.16

FACTORY LIGHTING:

Adequate lighting of factories is of vital importance, as it provides improved amenities for the employees, increased production and has a definite economic value in reducing accidents with their consequent loss of time and compensation payments.

General Requirements and Types of Installations: A factory lighting installation, in common with other indoor equipments should provide an adequate illumination on the working plane and give a good distribution of light, employ simple and easily cleaned fittings and avoid glare. It is essential not only to avoid glare from the lamp itself but also reflected glare from any polished surface, which may be within the line of vision.

General Lighting: The usual scheme in factories and workshops is to mount a number of lamps at a sufficient height so that uniform distribution of light over the working plane is obtained. In large machine shops the height is governed by the necessity of keeping the lamps above the travelling crane. In such cases it is often desirable to supplement the main lighting by side lighting in order to give additional illumination on a vertical plane. Since light coloured walls and ceiling add to the effectiveness of an installation, therefore; it is necessary to get whitewashing or painting done.

Local Lighting: On some points fairly intense illumination is required. For this purpose local lighting can be provided by means of adjustable fittings attached to the machine or bench in question or mounted on portable floor standards. Such lamps should be mounted in deep reflectors so that glare is avoided.

Portable hand lamps attached to wall plugs by means of trailing leads are used for maintenance work and emergency lighting. Low voltage lamps of not more than 50 volts are recommended for use as portable hand lamps because such lamps have thicker filament and are, therefore, more robust than those for normal voltage and danger of shock is also avoided in these few volt lamps. The supply for such lamps can be obtained from a special low voltage distribution system running throughout the factory or by means of a small transformer for each individual lamp.

Local lighting should never be employed alone, good general lighting is essential so that the dark places between the local lighting units are avoided. Dark places between the local lighting units cause fatigue to the eyes on account of its continually having to adjust itself to new conditions.

Emergency Lighting: Some lights, such as for *(i)* internal pilot lighting required for safe and speedy evacuation of personnel after main lighting circuit is off *(ii)* external pilot lighting, provided with careful shades leading to shelters required for evacuation of personnel *(iii)* for control posts, first aid centres etc. *(iv)* dials and gauges in important plants required to be watched regularly are required during an air raid when all the factory lights are off as a matter of air-raid precaution. The circuit supplying the above emergency lights should be independently controlled. It is very desirable to provide auxiliary lighting from the source other than the main electric supply preferably from batteries or from small petrol driven generator set. If, however, emergency lighting circuits are operated from main electric supply, these should be completely separated from main lighting circuit.

Industrial Lighting Fittings

Reflectors for industrial purposes must be simple in design and easily cleaned. The requirements of most of the installations can be met by one of the following types of fittings.

Standard Reflectors: These reflectors are made to accommodate lamps of ratings from 40 to 1,500 watts and designed so that they give adequate and uniform illumination when they are mounted at a spacing equal to about 1.5 times their mounting height above the working plane.

Diffusing Fitting: When more diffused light is required than that given by the standard reflector, a diffusing glass screen may be fixed across a standard type of reflector. Such fittings are used where highly polished articles are dealt with.

Concentrating Reflectors: A reflector with a concentrated beam is employed in large machine shops and foundaries, where the fittings are to be mounted on a considerable height above the working plane. In such places an ordinary reflector would have too wide angle of divergence and would waste a great deal of light on the walls.

Enclosed Diffusing Fittings: An opal globe completely enclosing the lamp giving a very even and well diffused light is used when light coloured walls and ceiling are there.

Angle Reflectors: Angle reflectors are used to provide illumination in a vertical plane where concentrating type reflectors are used. These can be mounted on suitable stanchions or the walls.

Maintenance

In order to maintain the fittings in a condition of reasonable efficiency it is necessary to clean the light fittings periodically. The frequency of cleaning depends on the conditions in the particular factory under consideration and varies from once or twice a week for very dirty surroundings to every four or six weeks under the best conditions.

Types of Lamps

The discharge lamps have been used in where colour rendering is not important. The fluorescent lamps are widely employed on account of its natural day-light colour, it's even illumination and absence of glare and in some cases, the fact that it gives rise to considerably less heat than filament lamps of the same light output.

ARTIFICIAL SOURCES OF LIGHT:

According to principle of operation the light sources may be grouped as follows:

1. **Arc Lamps:** Electric discharge through air provides intense light. This principle is utilized in arc lamps.
2. **High Temperature Lamps:** Oil and gas lamps and incandescent filament type lamps, which emit light when heated to high temperature.
3. **Gaseous Discharge Lamps:** Under certain conditions, it is possible to pass electric current through a gas or metal vapour, which is accompanied by visible radiations. Sodium and mercury vapour lamps operate on this principle.

4. **Fluorescent Type Lamps:** Certain materials, when exposed to ultraviolet rays, transform the absorbed energy into radiations of longer wave length lying within the visible range. This principle is employed in fluorescent lamps.

Incandescent Lamp:

An incandescent lamp essentially consists of a fine wire of a high-resistance metal placed in an evacuated glass bulb and heated to luminosity by the passage of current through it. Such lamps were produced commercially for the first time by Edison in 1879. His early lamps had filaments of carbonized paper which were, later on, replaced by carbonized bamboo. They had the disadvantage of negative temperature coefficient of resistivity. In 1905, the metallized carbon-filament lamps were put in the market whose filaments had a positive temperature coefficient of resistivity (like metals). Such lamps gave 4 lm/W.

At approximately the same time, osmium lamps were manufactured which had filaments made of osmium which is very rare and expensive metal. Such lamps had a very fair maintenance of candle-power during their useful life and an average efficiency of 5 lm/W. However, osmium filaments were found to be very fragile.

In 1906 tantalum lamps having filaments of tantalum were produced which had an initial efficiency of 5 lm/watt.

All these lamps were superseded by tungsten lamps which were commercially produced in about 1937 or so. The superiority of tungsten lies mainly in its ability to withstand a high operating temperature without undue vaporisation of the filament. The necessity of high working temperature is due to the fact that the amount of visible radiation increases with temperature and so does the radiant efficiency of the luminous source. The melting temperature of tungsten is 3655°K whereas that of osmium is 2972°K and that of tantalum is 3172°K. Actually, carbon has a higher melting point than tungsten but its operating temperature is limited to about 2073°K because of rapid vaporization beyond this temperature.

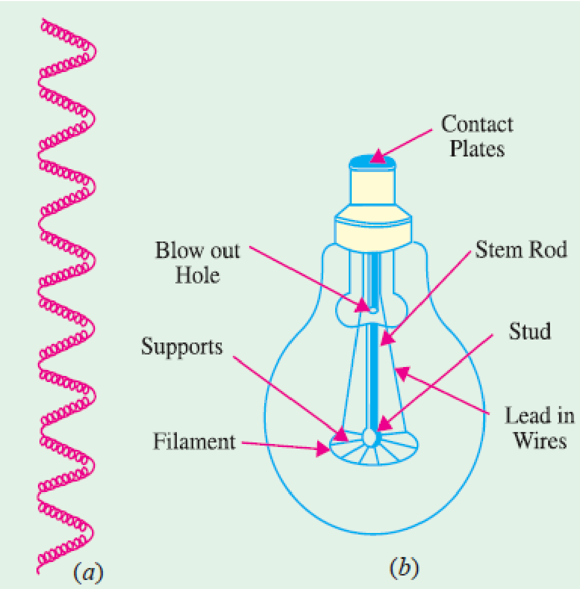
In fact, the ideal material for the filament of incandescent lamps is one which has the following properties :

1. A high melting and hence operating temperature
2. A low vapour pressure
3. A high specific resistance and a low temperature coefficient
4. Ductility and
5. Sufficient mechanical strength to withstand vibrations.

Since tungsten possesses practically all the above mentioned qualities, it is used in almost all modern incandescent lamps. The earlier lamps had a square-cage type filament supported from a central glass stem enclosed in an evacuated glass bulb. The object of vacuum was two fold : ***(a)*** to prevent oxidation and ***(b)*** to minimize loss of heat by convection and the consequent lowering of filament temperature.

However, vacuum favoured the evaporation of the filament with the resulting blackening of the lamp so that the operating temperature had to be kept as low as 2670º K with serious loss in luminous efficiency.

It was, later on, found that this difficulty could be solved to a great extent by inserting a chemically inert gas like nitrogen or argon. The presence of these gases within the glass bulb decreased the evaporation of the filament and so lengthened its life. The filament could now be run at a relatively higher temperature and hence higher luminous efficiency could be realized. In practice, it was found that an admixture of 85% argon and about 15 percent nitrogen gave the best results.



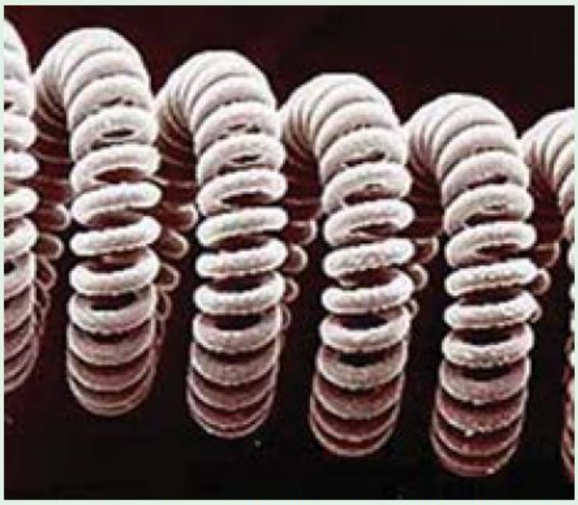

A electric filament is a metallic wire, usually made of tungsten, when heated to luminance by passing electricity, radiates light

Fig. 4.17

However, introduction of gas led to another difficulty i.e. loss of heat due to convection which offsets the additional increase in efficiency. However, it was found that for securing greater efficiency, a concentrated filament having a tightly-wound helical construction was necessary. Such a coiled filament was less exposed to circulating gases, its turns supplying heat to each other and further the filament was mechanically stronger. The latest improvement is that the coiled filament is itself 'coiled' resulting in 'coiled-coil' filament figure 4.17 (a) which leads to further concentrating the heat, reducing the effective exposure to gases and allows higher temperature operation, thus giving greater efficiency.

The construction of a modern coiled coil gas-filled filament lamp is shown in figure 4.17 (b). The lamp has a 'wreath' filament i.e. a coiled filament arranged in the form of a wreath on radial supports.

Fluorescent Lamp:

A **fluorescent lamp** or a **fluorescent tube** is a low weight mercury vapour lamp that uses fluorescence to deliver visible light. An electric current in the gas energizes mercury vapor which delivers ultraviolet radiation through discharge process which causes a phosphor coating of the lamp inner wall to radiate visible light. A fluorescent lamp changes over electrical vitality into useful light a great deal more proficiently than incandescent lamps. The normal luminous viability of fluorescent lighting frameworks is 50-100 lumens for every watt, a few times the adequacy of incandescent lamps with equivalent light yield.

Working of Fluorescent lamp:

- When the switch is ON, full voltage will come across the tube light through ballast and fluorescent lamp starter. No discharge happens initially i.e. no lumen output from the lamp.

- At that full voltage first the glow discharge is established in the starter. This is because the electrodes gap in the neon bulb of starter is much lesser than that of inside the fluorescent lamp.
- Then gas inside the starter gets ionized due to this full voltage and heats the bimetallic strip that is caused to be bent to connect to the fixed contact. Current starts flowing through the starter. Although the ionization potential of the neon is little bit more than that of the argon but still due to small electrode gap high voltage gradient is appeared in the neon bulb and hence glow discharge is started first in starter.
- As voltage gets reduced due to the current causes a voltage drop across the inductor, the strip cools and breaks away from the fixed contact. At that moment a large L di/dt voltage surge comes across the inductor at the time of breaking.
- This high valued surge comes across the tube light electrodes and strike penning mixture (mixture argon gas and mercury vapor).
- Gas discharge process continues and current gets path to flow through the tube light gas only due to low resistance as compared to resistance of starter.
- The discharge of mercury atoms produces ultraviolet radiation which in turn excites the phosphor powder coating to radiate visible light.
- Starter gets inactive during operation of tube light.

Construction of Fluorescent Lamp:

A **fluorescent tube light** consists of

1. a lime glass tube
2. drop of mercury
3. argon gas
4. phosphor coating
5. electrode coils
6. mounting assemblies
7. aluminum cap

Total set up of a lamp requires two bases and electromagnetic ballast or choke coil with a starter.

- The electrode mount assemblies are at both the ends of lamp tube.
- This electrode mounting assembly is almost similar to the stem press unit in the incandescent lamps.
- The electrode is similar to the incandescent lamp filament.
- The filaments of electrodes play both roles as anode and cathode.
- Small plates are attached to the filament to protect the electron bombardment and reduce the wattage loss at both ends.
- The filament is dipped in a mixture of barium, strontium and calcium carbonate. It is baked during manufacturing to become oxides and thus it becomes able to provide abundance of free electrons easily.
- Liquid mercury is provided inside the lamp bulb.
- Phosphor coating is used on inner wall of the bulb tube.
- At a certain pressure argon gas is filled up inside the tube.

- Two pins at each end are taken out of the lamp body through the cap.

The figure of an electrode is shown below.

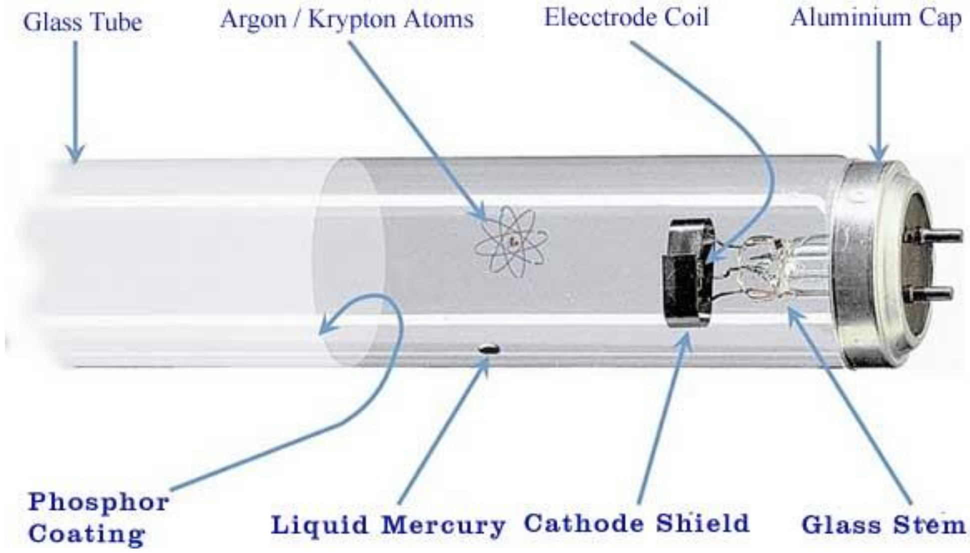


Fig. 4.18

A fluorescent lamp tube is loaded with a gas containing low-pressure mercury vapour and argon. The pressure inside the lamp is around 0.3% of environmental pressure. The inward surface of the lamp is coated with a fluorescent (and frequently marginally luminous). This coating is made of shifting mixes of metallic and uncommon earth phosphor salts. The lamp's anodes are normally made of snaked tungsten and typically alluded to as cathodes due to their prime capacity of discharging electrons. For this, they are coated with a blend of barium, strontium, and calcium oxides to have a low thermionic emanation temperature. Fluorescent lamp tubes are normally straight and long. The length of the commonly used lamp is around 100 millimetres (3.9 in). A few lamps have the tube twisted into a circle, utilized for table lamps or different spots where a more conservative light source is required. Bigger U-shaped lamps are utilized to give the same measure of light in a smaller region. Minimal fluorescent lamps have a few little width tubes joined in a heap of two, four, or six or a little breadth tube curled into a helix, to give a high measure of light yield in little volume.

To construct a fluorescent tube light a lime glass tube, drop of mercury, argon gas, phosphor coating and the electrodes with their mount assemblies are required. Total set up of a lamp requires two bases and choke coil with a starter. The electrode mount assembly is almost similar to the stem press unit in the incandescent lamps. The filaments play both roles as anode and cathode. Generally, small plates are attached to the filament to protect it from electron bombardment and to reduce the wattage loss at both ends. The electrode is similar to the incandescent lamp filament. But an exception is that this filament is dipped in a mixture of barium, strontium, and calcium carbonate. It is baked during manufacturing to become oxides and thus it becomes able to provide the abundance of free electrons easily.

Wiring Diagram for a Single Tube Light Circuit:

Tube light is not connected in the supply main directly. Although it operates at 230 V, 50 Hz, some auxiliary electrical components are used to insert in this installation to support the tube light operational principle. The total electrical components for single tube light installation are

1. Choke: it is electromagnetic ballast or electronic ballast.
2. Starter: Small neon glow up lamp
3. Switch
4. Wires

Wiring Diagram of Single Tube Light Installation with Electromagnetic Ballast:

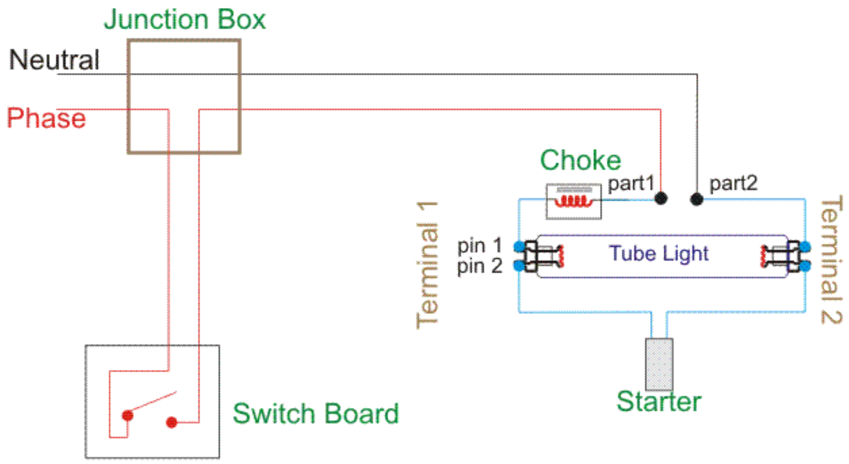


Fig. 4.19

Wiring Diagram of Single Tube Light Installation with Electronic Ballast:

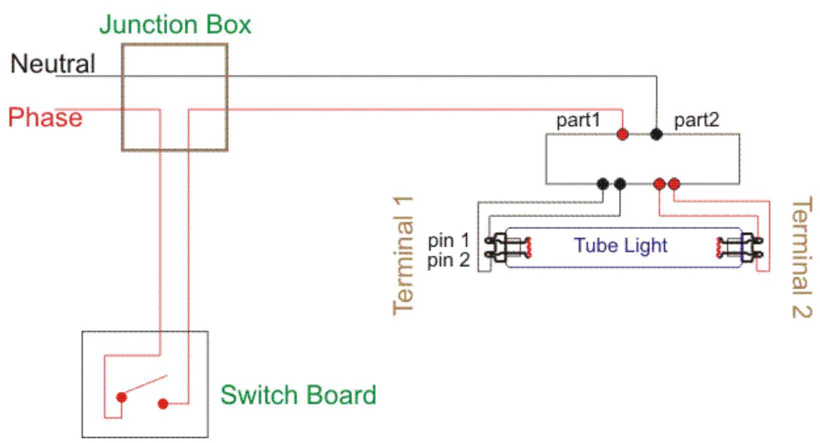


Fig. 4.20

Electronic Ballast:

Electronic ballast is a device that converts power frequency to very high frequency to initialize gas discharge process in Fluorescent Lamp by controlling voltage across the lamp and current through the lamp.

Use of Electronics Ballast

There are some advantages to use electronic ballast instead of electromagnetic ballast.

1. It operates in low supply voltage. It produces high frequency to give very high output voltage initially to start up the discharge process.
2. It creates very low noise during operation.

3. It does not create any stroboscopic effect or RF interference.
4. As it works with very high frequency, it helps to start the lamp operation instantaneously.
5. It does not require any starter that is use in electromagnetic ballast.
6. It never creates flickering.
7. No start up vibration occurs.
8. Its weight is very minimal.
9. Ballast loss is very less. Hence energy saving is possible.
10. It increases the life of the Lamp.
11. Due to operation at higher frequency, discharge process in fluorescent lamp is at higher rate. Hence quality of light is increased.

Working Principle of Electronic Ballast:

Electronic ballast takes supply at 50 – 60 Hz. It first converts AC voltage into DC voltage. After that, filtration of this DC voltage is done by using capacitor configuration. Now filtered DC voltage is fed to the high frequency oscillation stage where oscillation is typically square wave and frequency range is from 20 kHz to 80 kHz.

Hence output current is with very high frequency. A small amount of inductance is provided to be associated with high rate of change of current on high frequency to generate high valued $L\frac{dI}{dt}$. Generally more than 400 V is required to strike the gas discharge process in fluorescent tube light. When switch is ON, initial voltage across the lamp becomes 1000 V around due to high valued $L\frac{dI}{dt}$, hence gas discharge takes place instantaneously.

Once the discharge process is started, the voltage across lamp is decreased below 230V up to 125V and then this electronic ballast allows limited current to flow through this lamp. This control of voltage and current is done by control unit of the electronic ballast. In running condition of fluorescent lamp electronic ballast acts as a dimmer to limit current and voltage.

Basic Circuitry of Electronic Ballast:

In present days, electronic ballast design is so robust and somewhat complicated to work very smoothly with high leveled controlling ability. The basic components used in the Electronic Ballast are listed below.

1. **EMI filter:** It blocks Electromagnetic Interference if any.
2. **Rectifier:** It converts AC power to DC power.
3. **PFC:** It does Power Factor Correction
4. **Half Bridge Resonant Output:** It converts DC to square waved voltage with high frequency (20 kHz to 80 kHz).
5. **Control Circuit:** It controls voltage and current across and through the lamp respectively.

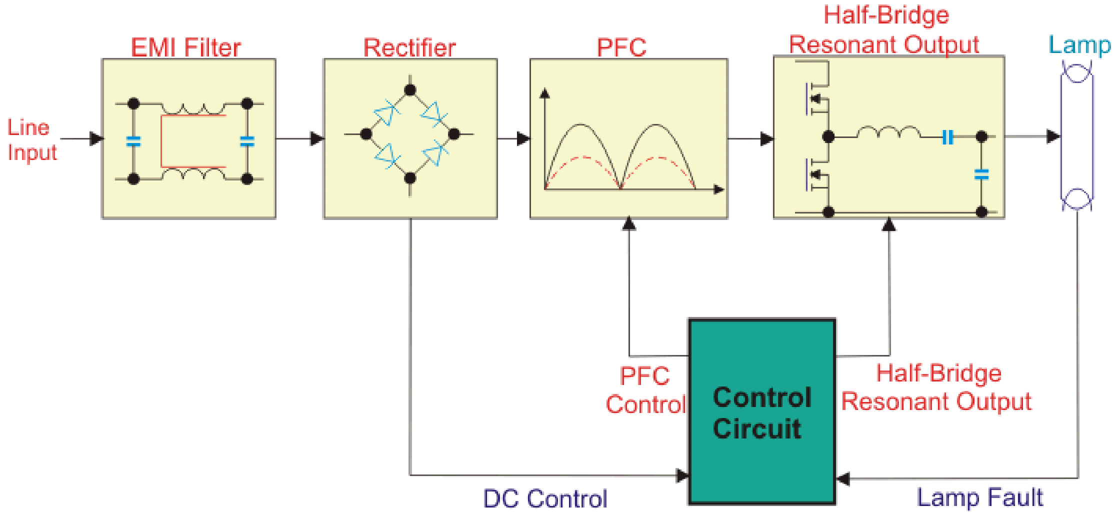


Fig. 4.21

Compact Fluorescent Lamp (CFL):

A compact fluorescent lamp (CFL), also known as a *fluorescent light* or energy saving light, is a type of fluorescent lamp which combines the energy efficiency of fluorescent lighting with the convenience and popularity of incandescent fixtures. CFLs can replace incandescents that are roughly 3-4 times their wattage, saving up to 75% of the initial light energy. Although CFLs costs 3-10 times more than comparable incandescent lamps, they last about 10 times as long (10,000 hours).

CFLs are most cost effective in areas where lights are ON for long periods of time.

Because of the potential of reducing electric consumption and pollution, some organizations have encouraged the adoption of CFLs. Some electric utilities and local governments have subsidized CFLs as a means of reducing electric demand. In India also almost all the states are encouraging the adoption of CFLs in place of incandescent lamps and are providing incentives or even distributing CFLs to people free of cost to save electricity.

CFLs radiate a light spectrum different from that of incandescent lamps. Improved phosphor formulations have improved the perceived colour of the light emitted by CFLs such that some sources rate the best “soft white” CFLs as subjectively similar in colour to standard incandescent lamps.

Like, all fluorescent lamps, CFLs contain mercury, which makes their disposal complicated.

Operation: CFLs work much like standard fluorescent lamps. They consist of two components: a gas-filled tube, and a magnetic or electronic ballast. The gas in the tube glows with ultraviolet light when switched ON and electric current from the ballast flows through it. This in turn excites a white phosphor coating on the inside of the tube, which emits visible light throughout the tube surface.

CFLs with magnetic ballasts flicker slightly at start. They are also heavier than those with electronic ballasts. This may make them too heavy for some light fixtures. Electronic ballasts are more expensive, but light immediately (especially at low temperature). They are also more efficient than magnetic ballasts. The tubes will last about 10,000 hours and the ballast about 50,000 hours. Most currently available CFLs have electronic ballasts.

Electronic ballasts contain a small circuit board with rectifiers, a filter capacitor and usually two switching transistors connected as a high-frequency resonant series dc to ac inverter, The resulting high frequency (about 40 kHz or higher) is applied to the lamp tube. Since the resonant converter tends to stabilize lamp current and so the light produced over a range of input voltages, standard CFLs do not respond well in dimming applications and special lamps are required for dimming services.

CFLs are designed to operate within a specified temperature range. Temperatures below the range cause reduced output. Most of the CFLs are for indoor use, but there are models available for outdoor use. Outdoor CFLs need installation in enclosed fixtures so as to minimize the adverse effects of low temperatures.

Types of CFLs: CFLs are available in a variety of styles or shapes. Some have two, four, or six tubes while others have circular or spiral-shaped tubes. The size or total surface area of the tube(s) determines how much light it produces. Some CFLs have the tubes and ballast permanently connected. Other CFLs have separate tubes and ballasts which facilitates in replacement of tubes without changing the ballast. They are also of types enclosed in a glass tube. These look somewhat similar to conventional incandescent lamps, except they are large.

CFLs are of two types: integrated and non-integrated lamps.

Integrated lamps combine a tube, an electronic ballast and either an Edison screw or a bayonet in a single unit. The incandescent lamps can be easily replaced with CFLs. Integrated CFLs work well in many standard incandescent light fixtures, reducing the cost of conversion.

Non-integrated CFLs have the ballast permanently in the luminaire, and only lamp is usually replaced at the end of life. Since the ballasts are placed in the light fixture, they are larger and last longer in comparison to the integrated ones Non- integrated CFL housings can be both more expensive and sophisticated. They have two types of tubes: bi-pin tubes and quad-pin tubes. Bi-pin tubes are designed for conventional ballasts, and quad-pin tubes are designed for electronic ballasts or conventional ballasts with external starters. A bi-pin tube contains an integrated starter which obviates the need for external beating pins but causes incompatibility with electronic ballasts.

CFLs are produced for operation on both dc and ac supply. DC CFLs are popular for use in recreational vehicles and off- the-grid housing. CFLs can also be operated with solar powered street lights, using solar panels located on the top or sides of a pole and light fixtures that are specially wired to use the lamps.

Advantages of CFLs over Incandescent Lamps

1. *Life Span.* As already mentioned, the average rated life of a CFL is about 10 times that of incandescent lamps.

The lifetime of any lamp depends on various factors, such as operating voltage, exposure to voltage spikes, mechanical shock, frequency of cycling on and off, lamp orientation, ambient temperature, manufacturing quality etc. The life of a CFL is significantly reduced if it is turned on and off frequently. In the case of a 5 minute on/off cycle the life span of a CFL can be reduced to “close to that of incandescent lamp”.

CFLs provide less light in their lives than when they are new. The light output decay is exponential, with the fastest losses being soon after the lamp is first used. By the end of their lives, they can be expected to produce 70-80% of their original light output.

2. *Energy Efficiency.* For a given light output, CFLs consume 20 to 33 per cent of the power of equivalent incandescent lamps. Electrical power equivalents for different lamps are indicated below.

Electrical Power Consumption In Watts		Minimum Light Output la Lumens
CFLs	Incandescent Lamps	
8-9	40	450
9-15	60	800
15-20	75	1100
20-25	100	1600
25-45	150	2600

3. *Heating and Cooling.* If in a building, indoor incandescent lamps are replaced by CFLs, the heat produced due to lighting will be significantly reduced. In warm climates office/ industrial buildings where air-conditioning is usually required, CFLs would reduce the load on the cooling system resulting in saving in electrical consumption, in addition to the energy efficiency savings of using CFLs instead of incandescent lamps. However, in cooler climates in which buildings need heating, the heating system would need more electrical energy. While the CFLs are still saving electricity, total greenhouse gas emissions may increase in certain scenarios, such as the operation of natural gas furnace to replace unintended heating from CFLs running on low-GHG electricity. Thus overall energy cost saving depends on the climate, increased heating demand offsets some of the lighting energy saved.

4. *Efficiency.* The typical luminous efficiency of CFLs is I 60 to 72 lumens/watt, and that of normal domestic incandescent lamps is 13 to 18 lumens/watt. Compared to a theoretical 100% efficient lamp, these figures are equivalent to lighting efficiency ranges of 9 to 11% for CFLs (60/680 to 72/680), and 1.9 to 2.6% for incandescent (13/680 to 18/680).

5. *Embodied Energy.* No doubt, CFLs need more energy in manufacturing than the incandescent lamp but this embodied energy is offset by their longer life and small energy consumption than equivalent incandescent lamps.

6. *Cost.* While the purchase price of an integrated CFL is typically 3 to 10 times higher than that of an equivalent incandescent lamp, the extended lifetime and lower energy consumption will more than compensate for the higher initial cost. CFLs are extremely cost-effective in commercial buildings when used to replace incandescent lamps.

7. *Starting Time.* Incandescent lamps attain full brilliancy a fraction of second after being switched on, CFLs turn on within a second, but many still take time to warm up to full

brightness. The light colour may be slightly different immediately after being switched on. Some CFLs are marketed as “instant on" and have no noticeable warm-up period, but others can take up to a minute to attain full brightness, or longer in very cold temperatures. Some that make use of a mercury amalgam can take up to three minutes to attain full brightness. This and shorter life of CFLs when turned on and off for short periods may make CFLs less suitable for applications like motion-activated lighting.

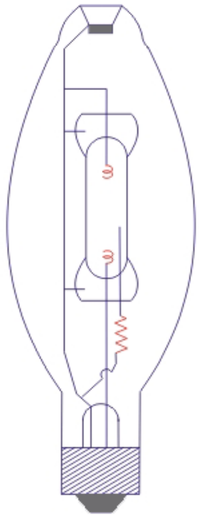
Disadvantages of CFLs

1. *Health Issues.* If individuals are exposed to the light produced by some single-envelope CFLs for long periods of time at distances of less than 20 cm. it could lead to ultraviolet exposures approaching the current workplace limit set to protect workers from skin and retinal damage. The ultraviolet radiation received from CFLs is too small to contribute to skin cancer and the use of double-envelope CFL lamps “largely or entirely” mitigates any other risks.
2. *Environmental Issues.* CFLs, like all fluorescent lamps, contain small amounts of mercury as vapour inside the glass tubing. Because mercury is poisonous, even these small amounts are a concern for landfills and waste incinerators where the mercury from lamps may be released and contribute to air and water pollution. Health and environmental concerns about mercury have promoted many jursdictions to require used lamps to be properly disposed or recycled rather than being included in the general waste stream sent to land fills. So disposal of used CFLs required proper care.

Mercury Vapour Lamp:

In case of fluorescent lamp the **mercury vapour** pressure is maintained at lower level such that 60% of the total input energy gets converted into 253.7 nm single line. Again transition of the electrons requires least amount of input energy from a colliding electron. As pressure increases the chance of multiple collisions gets increased. A schematic diagram of mercury lamp is shown below. This lamp is containing an inner quartz arc tube and outer borosilicate glass envelope. The quartz tube is able to withstand arc temperature 1300oK, whereas the outer tube withstands only 700oK.

Between two tubes nitrogen gas is used to be filled to provide thermal insulation. This insulation is for to protect the metal parts from oxidation due to higher arc temperature. The arc tube contains the mercury and argon gas. Its operational function is same as the fluorescent lamp. Two main electrodes and a starting electrode are inside the arc tube. Each main electrode holds a tungsten rod and upon which a double layer of coiled tungsten wire is wound. Basically the electrodes are dipped into a mixture of thorium, calcium and barium carbonates.



They are heated to convert these compounds into oxides after dipping. Thus they get thermally and chemically stable to produce electrons. The electrodes are connected through a quartz tube by molybdenum foil leads. Just when the main supply voltage is applied to the mercury lamp, this voltage comes across the starting electrode and the adjacent main electrode (bottom electrode) as well as across two main electrodes (bottom and top electrodes). As the gap between starting electrode and bottom main electrode is small the voltage gradient is high in this gap.

Because of this high voltage gradient across the stating electrode and the adjacent main electrode (bottom) a local argon arc is created, but the current gets limited by using a starting resistor. This initial arc heats up the mercury and vaporizes it and this mercury vapour helps to strike the main arc soon. But the resistance for the main arc current control resistor is somewhat less than the resistance of the resistor used in the initial arc current control purpose. For this reason initial arc stops and main arc continues to operate. It takes 5 to 7 minutes to make all of the mercury to be vaporized completely. The lamp gets its state of its operational stability. The mercury vapour arc gives visible spectra of green, yellow and violet. But there may be still some invisible ultraviolet radiation during discharging process of **mercury vapour** so phosphor coating may be provided on outer glass cover to improve efficiency of the mercury lamp.

There are five lamps with phosphor coating to provide improved colour performance. As the wattage increases the initial lumen ratings for the phosphor coated lamps get available with 4200, 8600, 12100, 22500 and 63000. The average life of mercury lamp is 24000 hours i.e. 2 years 8 months.

Comparison between Tungsten Filament Lamps and Fluorescent Tubes:

	Tungsten Filament Lamps	Fluorescent Tubes
1	Voltage fluctuation has comparatively more effect on the light output.	Voltage fluctuation has comparatively low effect on light output as the variations in voltage are absorbed in the choke.
2	Luminous efficiency increases with the increase in voltage of the lamp.	Luminous efficiency in creases with the increase in wattage and increase in length of tube.
3	It gives light close to natural light, therefore, objects are properly seen	It does not give light close to natural light; therefore, colour rendering is defective.
4	Luminous efficiency of coloured filament lamps is poor because coloured glasses are used for this purpose.	Different colour lights can be obtained by using different composition of fluorescent powder. Hence efficiency is high and better colours are obtained.
5	Due to comparatively high working temperature heat radiations are also present.	Due to low working temperature heat radiation is low.
6	Its brightness is more.	Its brightness is less.
7	With the time light output is reduced.	With the time light output is gradually reduced.
8	No stroboscopic effect.	It has objectionable stroboscopic effect.
9	Though the life of filament lamps varies with the working voltage, however, its normal life is 1,000 working hours.	Life of fluorescent tubes is not affected so much by variations in voltage but it depends on the frequency of starting. The life of the tube is about 7,500 working hours.
10	The initial cost per lamp is quite low.	The initial cost per tube is more.
11	For same lumens output more lamps are required and wiring cost is more. Life of lamp is also low. Hence overall cost of maintenance is more.	For same lumens output, lesser number of tubes is required, the wiring cost is low. Since its life is comparatively more, replacement cost is low. Hence overall maintenance cost is low.

Glare:

Glare is difficulty seeing in the presence of bright light such as direct or reflected sunlight or artificial light such as car headlamps at night. Because of this, some cars include mirrors with automatic anti-glare functions.

Glare is caused by a significant ratio of luminance between the task (that which is being looked at) and the glare source. Factors such as the angle between the task and the glare source and eye adaptation have significant impacts on the experience of glare.

Discomfort and Disability:

Glare can be generally divided into two types, discomfort glare and disability glare. Discomfort glare results in an instinctive desire to look away from a bright light source or difficulty in seeing a task. Disability glare impairs the vision of objects without necessarily causing discomfort. This could arise for instance when driving westward at sunset.

Disability glare is often caused by the inter-reflection of light within the eyeball, reducing the contrast between task and glare source to the point where the task cannot be distinguished. When glare is so intense that vision is completely impaired, it is sometimes called dazzle.

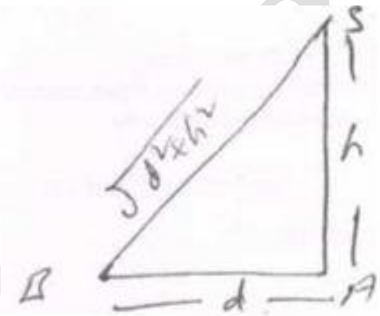
Glare Reduction:

When any light is misdirected and that usually shines directly on the eye instead of on the object we wish to see, it can be considered as glare. Glare can be reduced to acceptable levels by using several techniques, which keep this unwanted light out of the eye.

1) Use a large number of small sources rather than a small number of high-luminance sources.
2) Mount luminaries out of the field of view of the common work place.
3) Screening, shielding the sources from direct view or covering with diffusing plates or filters or cross polarizers greatly reduces glare.
4) Since glare is also due to sudden changes in illuminance, more uniform lighting can reduce it.
5) Educate workers not to shine cap lamps into other workers' eyes when travelling in cages or passing instructions.
6) Use proper or correct lighting and avoid specular materials such as metallic paint on mechanics or wall rock and choose a flat paint when possible.
7) Keep the surrounding and surrounding luminances high. Only 1/30th of the task luminance is necessary for accomplishing the task.

Questions Bank

1. State and explain two laws of illumination.
2. Describe the general principles that are employed in the design of street lighting installation.
3. Explain the direct lighting scheme and indirect lighting schemes.
4. Discuss the factory lighting.
5. With a neat diagram, explain the construction and working of the sodium vapour lamp.
6. Explain the working of a high pressure mercury vapour lamp with the help of a circuit diagram.
7. Discuss the requirements of good lighting.
8. Briefly explain i. Flood lighting. ii. Factory lighting with the requirements to be met in each case.
9. Define the following terms: i) Luminous flux ii) MHCP iii) Solid angle iv) MSCP
10. Write a note on Flood Lighting.
11. With a neat Sketch, explain the working principle and constructional details of a fluorescent Tube light**.**
12. Explain with a sketch the construction and working of compact fluorescent lamp.
13. Find the height at which a light source having uniform spherical distribution should be placed over a floor in order that the intensity of horizontal illumination at a given distance from its at a given distance from its vertical line may be greatest.



14. An incandescent lamp has suspended from the ceiling of a room. The illumination below the lamp vertically downwards is 80 lux. When the illumination is measured at a distance of 2m from the ceiling, its value is 40 lux. Find the candle power of the lamp and its vertical distance from the floor.
15. Two lamps A and B of 200 candela and 400 candela respectively are situated 100 m apart. The height of A above the ground level is 10 m and that of B is 20 m. If a photometer is placed at the centre of the line joining the two lamp posts, calculate its reading.
16. Define:
    (i) Luminous Efficiency, ii) Depreciation factor, iii) coefficient of utilisation
    iv) Space –Height ratio.
17. Define the following terms related to lighting.
    i. Light. ii. Luminous flux. iii. Luminous Intensity.
    iv. Lumen. v. Candle power.
18. What are the main objectives of street lighting and explain the principle to be employed

in discharging street lighting.

19. The illumination in a drawing office 30 m × 10 m is to have a value of 250 lux and is to be provided by a number of 300-W filament lamps. If the coefficient of utilization is 0.4 and the depreciation factor 0.9, determine the number of lamps required. The luminous efficiency of each lamp is 14 lm/W.
20. Two similar lamps having uniform intensity of 500CP in all directions below the horizontal are mounted at a height of 4m. What must be the maximum spacing between the lamps, so that the illumination on the ground mid way between the lamps shall be at least half the illumination directly under the lamps?
21. It is required to provide an illumination of 100 lumen/m^2 in a factory hall 40m x l0m x 5m. taking depreciation factor as 0.8, coefficient of utilization as 0.4, efficiency of lamp as 141u/watt, calculate the number of lamps and their arrangement.
22. A light source having an intensity of 200 C.P in all directions is fitted with a reflector so that it directs 85% of its light along a beam having a divergence 20°. Calculate:
    a) The total height flux emitted along the beam.
    b) The average illumination produced in a surface normal to the beam direction at a distance of 5 inches.
23. A building frontage 50mxl5m is to be illuminated by flood lighting projectors situated 25 meters away. If the illumination is 100 lux, coefficient of utilization 0.5, depreciation factor 1.5, waste light factor 1.2, estimate the number and size of projectors. Sketch the projector recommended indicating the usual adjustments provided.
24. The following data related to the lighting scheme of an engineering college:

    Dimensions of the hall = 30m x20mx8m (high)

    Mounting height = 5 m.

    Required level of illumination = 140 lumen/ m^2.

    Utilisation factor = 0.6.

    Maintenance factor = 0.75.

    Space-height ratio = 1.

    Lumens/ watt for 300 w lamp =13.

    Lumens/ watt for 500 w lamp =16.

    Estimate the number and size of single lamp luminaries with their disposition.