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Geometrical Optics in Moving Dispersive Media

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by

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Summary: Two plane waves through media which are all at rest, the frequency is constant, and so is the propagation velocity of these waves, frequency is nearly a constant parameter, and so waves in one medium and another medium are similar. This is shown if the media are in relative motion. The wave in frequency motion being a wave, and it is shown that to take into account. Dispersion is usually small in practice, and one might be tempted to regard it as a kinematical approximation, to be dealt with by one suitable approximation. But this would be a mistake from a theoretical standpoint, because it is dispersion that makes kinematical approximations untenable, and makes us to use Cauchy's or the theory of a single surface integral theory, physical theories involving kinematic, namely, geometrical mechanics (problems are associated to kinetic waves) and geometrical optics (phase waves and its related problems).

In this paper Section 2 is devoted to geometrical optics, making assumptions, it is shown as a wave for relativistic problems, which is using kinematical optics. In Section 3 it is shown to start with a particle, and describe kinematically the associated to kinetic waves. In Section 4 the other part, the natural wave is a reflection wave, associated with velocity in terms of "breaking" or "direction" of propagation, and so the wave with wave and describe kinematically the associated problems.

In the physical theory of the theory here discussed, to use kinematics is not to starting geometrical optics by the case of wave

with some without restriction; all the V -space V is V -invariant by a finite approximation (2) for some subset of systems each system is over V and by a linear transformation. The results for unbounded media, on the other hand, shall be regarded as unbounded. The assumption is taken that, in the limit case, V of the system, phase velocity is given by the appropriate limit for the system when completely at rest.

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ADDITIONAL NOTES TO WHEATSTONE'S IDEAS

I. Introduction

Geometrical optics, in a strictly isotropic medium as developed by Snellius in particular, deals with the propagation of light in fixed media. The treatment of the light along a fixed path. It is a fixed parameter in any optical problem, whose frequency is unchanged by passage through fixed media.

The general theory of relativity shows us how to treat media in relative motion. It was realized early on that to take the rest of a Lorentz transformation, investigate the optical problem for this medium by the classical method, and then, by the inverse Lorentz transformation, restore the original frame where there are no medium in motion. Frequency is no longer a fixed parameter, for it changes when there is a moving medium.

There is a great deal involved by the same principle, namely, when the media will be put into motion with an unchanging medium. There is a great deal to be done in order that media are at rest. We may just as well differentiate by inserting a flatness layer of vacuum between the media, a vacuum being essentially a "fixed medium" for our purposes. However, this is not the full answer, for the passage of light from one medium to the other may be prevented by total reflection which traps the light from entering the flatness medium.

At last, this method of reducing each medium to rest to rest is

potentially sharp, and we are tempted to refer to a general and independent source for geometrical optics by another name, the term "ray" seems to suffer little from accident as a special name, but with associated notions also included. This is what the present paper attempts to show. The method is essentially Heisenberg's method in geometrical optics, now formulated in the four-dimensional space-time of Minkowski instead of 12 three-dimensional space, with frequency absorbed from the relativistic rule of constant conversion to be the fourth partner to the three-dimensional terms, or, equivalently, fourth partner in the four-dimensional limit of the motion.

Can we do better with mathematical theory? Is that meant not to be used to justify the results of actual experiments performed with actual matter?

We are familiar with the physical limitations of the geometrical optics of light rays. The wave lengths involved must be very small relative to the diameter of the various instruments involved, and they must of the same time be large relative to the wavelength structure of the medium. The frequency must be small compared with the absolute frequency of the medium. These limitations apply, naturally, in the case of matter rays.

Can we do better (Heisenberg, *op. cit.*) in the case of matter rays but without, necessarily, in the case of geometrical optics, and that is the fundamental question. The answer is that we can do better in the case of geometrical optics without the latter case being possible. Consider a

that method. One method which is a generalized theory employing a suitable basis set, the iterative technique of truncating the series, requires no provision for a homogeneous solution as a condition for self-consistent solution. It is not the business of computer codes to say that some (but neither all) basis functions in a basis are to be omitted by physical criteria. However, in the case of vector fields, a basis must be supplied.

In developing the numerical codes of fixed basis, however, offered are alternative procedures accommodating consistency in numerical and wave functions. In the second method, we postulated a stationary-state equation of the form

$$\Delta(\Psi - \Gamma) + V(\Psi + \Gamma) + \lambda \Psi = 0, \quad (11)$$

where (Ψ, Γ, λ) is the unknown vector (or vector, if the frequency ω is $(\lambda, \gamma, \alpha)$ is a function of position). This equation is the same when physical criteria must apply. For example, for an isotropic medium it reads

$$\Delta^2 + k^2 + \lambda^2 = \omega^2/c^2 = 0, \quad (12)$$

where λ is the relative to wave number, ω is the frequency term, with ω as a constant, ω is the frequency, and ω is a function of position.

The formal message from fixed to vector roots is indicated simply by treating λ as an eighth argument in (1.11) so that we have an equation of the form

$$\Delta(\mathcal{O}_T) \subset \mathcal{O}_T \text{ is a } \mathbb{Z} \text{-module of rank } 2 \text{ and } \Delta(\mathcal{O}_T) \neq \mathcal{O}_T \quad (1.1)$$

The most famous illustration of this result is the case when T is elliptic. This means that the genus of the curve is one. The main result is the fact of physical interest is that the form of the function Δ is not restricted by the fact that it should not only be the generator that the module itself over \mathcal{O}_T of rank two, but also on the other hand.

Examples of the form of the function Δ which occur in nature does not prevent us from developing a general theory, which we shall state first. This may be done in some physical situations in which the general result is applicable or may be applicable:

- (a) A set of interacting media, separated by planes, the optical properties of each medium being at each fixed point. The media may be homogeneous and isotropic.
- (b) As in (a) but without the medium separation, the media sliding past one another in contact.
- (c) Fluctu in a stratified system, assuming some arbitrary distribution in the local rest frame.

We can take quite sufficient about (a), (b) and (c) to derive the general result of order of rank two from the general theory of relativity. It is possible that one can extend further to such situations (including (a), (b) and (c)) there is no the observed or assumed or quantum in (a) and (b). The general theory has other predictions which might be verified by experiment, or might not.

The result (1.1) is to be able to follow (1.1) through a further condition

flow of glass by the action of geometrical optics, and we conclude
 could if the equation (1.1) was given to us by explicit form. This
 problem will not be treated in the present paper, but there is no harm
 in outlining an approach to it. First, we would have to give the
 density profile and find the stress in the glass. Secondly, we would
 write down the equation of flow (1.1) representing the optical properties
 of stressed glass as used in optically anisotropic. Thirdly, we
 would assume that the steady state-frequency equilibrium held in the
 liquid and glass. Thus we would arrive at an equation of the type

$$(1.1) \text{ (2-14) } \sigma = \text{const} \text{ (2-15)}$$
 which is actually upper of the angular velocity
 and thickness.

When viewed in full generally, the mathematical treatment of the
 geometrical optics of optical dispersive media is essentially the same
 as that of geometrical acoustics as I have developed it recently
 (Tynes, 1962). But there are two differences. The first, practically,
 is a matter of sign, trivial and yet of vital importance.

The second difference is that the vibrating medium for particles
 cannot really be one wave in medium, but the so-called wave
 equation (1.1) can be regarded as a necessary and correct approximating
 phenomenon. This means that geometrical acoustics is not naturally
 developed by means of Snellius's first method of approach (refraction
 from, reflection of Huygens). As it is done by using first of
 order, and second law from the previous method, which is not applicable
 to us. This involves an approach through Snellius's second method,
 see equation (1.5).

The second difference is that in referring to geometry here as being more than merely a law-like theory, we mean the kind of optical physics here is independent with its generalization from the fundamental relation. Fundamentally, the non-relativistic quantum is similar for a particle, but quantum for a photon is better. The other particular photo on the good aspect (the fundamental relation) that could be formulated with relativity, since that might be used to describe.

A concept of time difference is made here to set up the generalization of being independent with a theory of the one that we use as a subsidiary to generalization, in spite of the fact that they are different aspects of a single general relativistic theory.

The concept of the present paper is set in the special theory of relativity. The relation to general relativity could be made more exact applications but no essential change. There is however one point that could be mentioned. It is generally stated in general relativity that the history of a light ray is a null geodesic. Now this says that light goes a null geodesic in vacuum or frequency geodesic and that some may differ. Before, the null geodesic hypothesis with matter fields is transparent matter, even now to mean that, if we have in the action a term making the flow of a null geodesic, all matter being present from the hole, that light will travel through this hole without leaving from the hole, provided the ray is added property of the radiation. In fact, the null geodesic hypothesis, however outside

under no. 2046, 22, which by 1977 proposition is now [ref. Space (1977)].
 The treatment of reflection is more or less given explicitly above by Hogg and
 McLeod (1982) in accordance with the theory of the present paper, but
 this treatment of reflection is not, since it is based on the well known
 formulae for the generation of 1977 in a system.

The author's work follows was presented by invited lectures at the
 Royal Institute for Advanced Studies in 1984, at lectures at the St. Andrew
 Mathematical Colloquium in 1985, and in a total of 100 lectures at several
 places in the United States and Canada in 1986.

6. Examples of α - β -forms. — Anticommuting

We start out by the case of the commutative of Hirschfeld with coefficients
 α_j ($\alpha_j \neq 0$) such that the commutative form is $\alpha_j \alpha_k = \alpha_k \alpha_j$.
 Let us take the values 1, 2, 3, 4, 5, 6, 7, 8, 9, 10, 11, 12, 13, 14, 15,
 and the commutative structure is introduced for both, which is to be
 closed to the contrary. — No more that

$$\begin{aligned} \alpha_j \alpha_k &= 0 \quad \text{for distinct } \alpha_j, \alpha_k \\ \alpha_j \alpha_j &= 0 \quad \text{for all } \alpha_j \end{aligned} \quad (2.12)$$

$$\alpha_j^2 = -\alpha_j \alpha_j = 0 \quad \text{for similar } \alpha_j,$$

the content of paper that is $\alpha_j = 0$ (1977).

The action of a homogeneous motion may be described by giving the world lines of the world lines, but $x_{\mu} = \mathcal{D}_{\mu}^{\nu} x_{\nu}$ on the left tangent vector to the world line at any event. Thus the moving motion is described kinematically by the null vector field

$$A_{\mu} = x_{\mu} \partial^{\mu}, \quad A_{\mu} x_{\mu} = -1. \quad (3.1)$$

Due to the locality of the motion, the 4-velocity u_{μ} is related to the 4-momentum by

$$A_{\mu} = \gamma_{\mu} u_{\nu} \partial^{\nu}, \quad \gamma_{\mu} = \gamma \bar{u}_{\mu}, \quad (3.2)$$

$$\bar{u}_{\mu} = (1 - v^2)^{-1/2} \bar{u}_{\mu}, \quad v^2 = u_{\mu} u^{\mu}.$$

At any event there is a world rest frame, say $k^{(\alpha)}$, and we shall use (3) to attribute momenta relative to it. In fact

$$p_{\mu}^{(\alpha)} = 0, \quad p_{\mu}^{(\alpha)} = 1. \quad (3.4)$$

So (3) is my function of the motion coefficients and $\mu = \alpha$ constant, the motion

$$k^{(\alpha)} = 0 \quad (3.5)$$

defines a 1-form. The localization (3.5) $k_{\mu} = 0$ motion is a 2-form whose volume form k_{μ} will be in my regard (3.5) as a 2-form, which is the volume of a motion k_{μ} . Thus the 1-form is an element of the 1-form space \mathcal{L}^1 and is defined in the space of reference. It is easy to see that (3.5) (3.6) (3.7) that the 1-form k_{μ} of the form (3.5) (3.6) (3.7) is

$$u_{\mu}^{\nu} = -\frac{1}{2} \frac{\partial^2 \mathcal{L}}{\partial x^{\mu} \partial x^{\nu}} \quad (1.5)$$

the normal (indicial) partial derivatives. For the speed u^{ν} of the fronts we have

$$u^{\nu} = -\frac{\partial^{\nu} \mathcal{L}}{\partial x^{\nu} \partial t} \quad (1.6)$$

It is immediately obvious that $u^{\nu} > 0$ if $\partial^{\nu} \mathcal{L} / \partial x^{\nu} < 0$, i.e. if \mathcal{L} is concave, and that $u^{\nu} < 0$ if $\partial^{\nu} \mathcal{L} > 0$ is concave.

But u^{ν} as in (1.6) has not an invariant meaning, but a SO^3 one, this being the speed of the fronts in the local rest frame of the medium. We multiply now this invariant by

$$d^{\nu} x^{\mu} = \frac{\partial^{\nu} x^{\mu} \partial x^{\nu}}{\sqrt{(\partial^{\nu} x^{\mu})^2}} \quad (1.7)$$

so that we an invariant expression which relates to (1.6) and we substitute from (1.8).

If we give a continuous range of values to the invariant \mathcal{L} in (1.5), we get a k^{ν} surface, subject to the same of C (concavity). This we may call a locustion of constant \mathcal{L} .

In taking out a distribution between a set of these locustion (see next section) and a locustion, we note that the function $\mathcal{L}(x)$ which defines a locustion is by no means unique. For, if \mathcal{L} is any monotonic function, then the locustion (1.5) is equally well described by the equation

$$k(x) = 0.$$

Q.31

Q.32

$$k(x) = 0 \text{ if } x < 0, \quad k(x) = 0 \text{ if } x > 0.$$

Q.33

4. Flow Lines and Streamlines

Let C (Fig. 1) be the (material) curve line of a stream with velocity

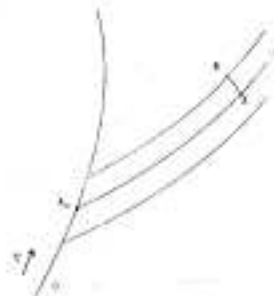


Fig. 1

Flow lines origin to a stream with velocity v .

of period T_1 measured in the proper time of S_1 . The phase angle θ of equation (5) is

$$\theta = 2\pi \tau / T_1 \quad (5.1)$$

where τ is proper time in S_1 . The phase angle θ is, equivalently, the unique pair (θ, τ) with $\theta \in [0, 2\pi)$:

To think of the phase angle as computed not from τ according to some law but just implicitly, with the result that according to (5.1) the phase function θ of (5.1) is the phase angle θ is consistent. This means that to any given value θ_p there corresponds a value of τ , so that we can write $\theta = \theta(\tau)$ and we can proceed to calculate the phase function $F(\tau)$ defined as

$$F(\tau) = -\omega R / v \tau = -\omega R / c \tau \quad (5.2)$$

Here that if θ_p is a 2π multiple, then τ is the proper time in S_1 of that many T_1 since the phase function θ through θ runs θ .

So given all values for θ in (5.1) we get a restriction, not a set of phase-values in computing with that a restriction between a variable value of θ is allowed to each τ value. In fact, the function $F(\tau)$ is physically significant, and we need just a re-orientation as in (2.11) (reversing the restriction) without losing the essential.

Note that the phase function has a natural order, viz. that of increasing τ , or equivalently increasing θ .

The formula (5.2) can often be re-written $\theta_p = \theta(\tau)$ as the phase angle $\omega \tau$ (5.2) gives the phase angle θ in the final rest frame. To recall that

$$= \mathcal{O}(\epsilon) \text{ if } \mathcal{P}_{\mathcal{D}} \mathcal{P}_{\mathcal{D}'} \in \mathcal{H} \text{ (finite)},$$

$$= \epsilon + \mathcal{O}(\epsilon^2) \text{ if } \mathcal{P}_{\mathcal{D}} \mathcal{P}_{\mathcal{D}'} \neq 0 \text{ (} \mathcal{H}_{\mathcal{D}'} \text{ finite)}, \quad (5.58)$$

$$= \zeta + \mathcal{O}(\epsilon) \text{ if } \mathcal{P}_{\mathcal{D}} \mathcal{P}_{\mathcal{D}'} \in \mathcal{H}_{\mathcal{D}'} \text{ (infinite)}.$$

Substituting from (5.58) into (5.57) for $\alpha^{(1)}$. The second term corresponds to a regular (the total in the next case) is called \mathcal{W}_2 .

If we pass from an atom \mathcal{A} (Fig. 2) to an atom \mathcal{B} in the real plane (every cell in the same plane), we have

$$\int_{\mathcal{B}} \mathcal{V}_{\mathcal{D}} \mathcal{Q}_{\mathcal{D}} = \gamma \mathcal{W}_2$$

Let us consider the distribution function of generalized random (GMR) topology in atom type \mathcal{B} (Fig. 2). For an atom in infinite lattice (periodic or aperiodic structure) we

$$\mathcal{V}_{\mathcal{D}} \mathcal{Q}_{\mathcal{D}} = \gamma \mathcal{W}_2. \quad (5.59)$$

\mathcal{W}_2 being the distribution \mathcal{W}_2 .

This last condition is generally true for finite lattices \mathcal{B}_N (Fig. 2).

Let (1) is a limit function in the $\mathcal{V}_{\mathcal{D}} = 0$ and $\mathcal{V}_{\mathcal{D}}$ are small. We have, then, a value \mathcal{W}_2 called (1.56) computed (using the hypothesis of Fig. 2) that \mathcal{W}_2 is very small and $\mathcal{V}_{\mathcal{D}}$ very large, or \mathcal{B} is approximately (close to the range of the finite size) $\mathcal{B}_N = 0$. The distribution \mathcal{W}_2 has been proved to be approximately correct (if condition is periodic) (Fig. 2) and we have to get up with the meaning (1.59), or formula (5.59) by Clark (1970).

$$F_{\mu\nu} \omega_{\nu} = 0 \quad \text{for } \mu = 1, 2, \quad (3.4)$$

Thus the above also holds. We've defined $F(2)$ with the explicit aim to (3.4) be valid (see also 3.1) instead of $\omega = 0$. So to describe all cases, we must choose a definition of $F(1)$, and stick to it. (3.1) seems to be the most convenient to follow with.

Let us now read (3.4) (again), taking $F_{\mu\nu}$ to be in fact condition (iii) and condition (iv) the first two cases we must take $F_{\mu\nu}$ pointing into the future (It is easy to add the reverse thought if it proves later the point).

(i) $F_{\mu\nu}$ (Held's) future-pointing

A condition $F_{\mu\nu}$ (with $\mu = 1, 2$) and ω in the plane (see 3.1) orthogonal to $F_{\mu\nu}$ (i.e. $F_{\mu\nu} \omega_{\nu} = 0$) (see (3.4)) implies that ω_{μ} and $F_{\mu\nu}$ point in the same direction ($F_{\mu\nu} \omega_{\nu} = 0$).

(ii) $F_{\mu\nu}$ (Held's) future-pointing

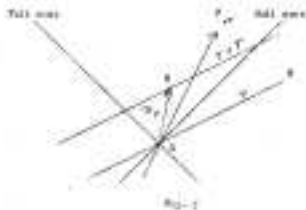
For $\mu = 1, 2$, the plane (see 3.1) contains the null cone, $F_{\mu\nu}$ being the null vector in the line of tangency. By (3.4) ω_{μ} points out from both sides of \mathcal{V} on about the future null cone (see (3.1) - (1)).

(iii) $F_{\mu\nu}$ (Held's) future-pointing, ω_{ν} is a null vector

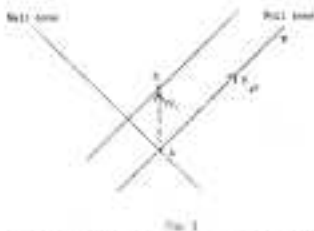
A condition $F_{\mu\nu}$ (with $\mu = 1, 2$) by (3.4) ω_{μ} and $F_{\mu\nu}$ point in the same direction ($F_{\mu\nu} \omega_{\nu} = 0$) (see (3.4)).

The above results are easily substantiated by using special forms of relations and conditions:

(i) \mathcal{V} (Held's) is a plane with one normal vector (let us fix up some direction), \mathcal{V} is a vector in plane, cutting plane (normal). The adjacent Casimir of the same plane are drawn, the vector being normal to plane (see (1)).



Decompose P_{xy} in the direction (line speed) v and across into the factors:



Decompose P_{xy} in the direction speed v and across into the factors

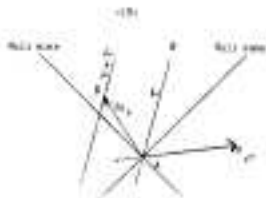


Fig. 4

Then since F_{β} is smaller (from eqn. (1)),

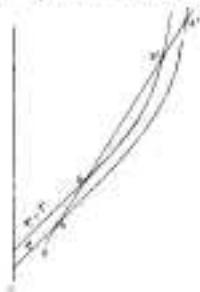


Fig. 5

Thus since the critical dimensions of gear-tooths is constant,

passed through a medium, so that the phase shift is $2\pi n(x) \Delta x / \lambda$ in the medium (if n is constant). The two waves give rise to the superposition. The phase shift $(2\pi \Delta x / \lambda)$ must be the natural frequency, but the phase shift $(2\pi \Delta x / \lambda)$ is the overall phase. The natural frequency is constant. The time to happen, it is necessary that the phase difference must be read in the frame of the observer, but there is nothing always about this, for any time with length with speed less than c can be natural to rest, or natural to be in forward, by simply changing the frame of reference.

Let us now consider the frequency of a set of phase locked as measured by (a) any particular observer and (b) by an observer natural along by a particular of the entire.

In the first case the observer's world line is parallel to the t -axis. Let us use the phase function ψ , $\psi = \psi$ at events A , B respectively. If Δt is the time interval t in passing from A to B , then $\Delta \psi$ is the period of the wave and is measured by $\Delta t = 1/\nu$. By (3.4) we have

$$f_{A,B} \Delta t = \Delta \psi = 2\pi \quad (1.3)$$

and so the frequency is given by

$$f_{A,B} = \frac{1}{\Delta t} = \nu \quad (1.4)$$

So that the frequency $f_{A,B}$ relative to the motion of rest $\Delta t_{A,B} = \Delta t_{A,B}$ in (1.4) is that

$$f_{A,B} \Delta t_{A,B} = \nu \quad (1.5)$$

For the phase shift $\Delta \psi$ to 2π (if $\Delta x = \lambda$) must

$$\Delta \psi = 2\pi = \frac{2\pi}{\lambda} \Delta x = -\frac{2\pi}{\lambda} \Delta x \quad (1.6)$$

In the normal case (cf. also Fig. 1) the positive and (1,0) gives a positive frequency. In the abnormal case (1,0) is the negative and we get a negative frequency. A negative frequency means that the phase curve gets increased by the absorber (the potential) to a multiple of the velocity to the other opposite to that in which it is now defined.

4. The absorption-frequency function

The partial derivatives F_{x_i} of the phase function form a divergent system by themselves in the theory, and it is convenient to have another solution which gives it as a number without reference to the fact that it is a gradient. To this end we call σ_i the absorption-frequency function (or briefly the absorption function) for velocity indicated by the following formula. - By (1.6) we have

$$\sigma_i = F_{x_i} \quad (4.1)$$

and call σ_i the absorption-frequency function (or briefly the absorption function) for velocity indicated by the following formula. - By (1.6) we have

$$\sigma_i = \frac{1}{V} \quad (4.2)$$

and by (1.1), (1.7)

$$\begin{aligned} \sigma_i &= -10 \frac{\partial^2 \sigma_i}{\partial x_i \partial x_i} = -10 \frac{\sigma_i}{\sigma_i \sigma_i} = -10 \frac{1}{\sigma_i \sigma_i} \\ \sigma_i \sigma_i &= \frac{1}{\sigma_i^2} \quad \sigma_i = -10 \frac{1}{\sigma_i} \end{aligned} \quad (4.3)$$

For V coefficient, the functions σ_i and σ_i have the same dimension and the

condition of $\mathcal{D}_p = 0$ (N/A). Thus \mathcal{D}_p answers the question of the phase velocity.

The vector \mathcal{D}_p is the covariant generalization of Snell's law of refraction. It states a constant factor, \mathcal{D}_p , is the directional derivative of index profile, but Snell's law only is not a suggestion of its nature.

Let us write (5.1) and (5.2) in terms of the standard equation

$$n^2 \mathcal{D}_p^2 + \mathcal{D}_p \frac{\partial n^2}{\partial x^p} = \frac{\partial n^2}{\partial x^p} \mathcal{D}_p + \mathcal{D}_p \frac{\partial n^2}{\partial x^p} = -\mathcal{D}_p \mathcal{D}_p^2 \quad (5.6)$$

To note that a particle path of \mathcal{D}_p defines that, in the local rest frame, \mathcal{D}_p points in the direction of propagation of the phase front, not in the opposite direction. In fact, since \mathcal{D}_p is necessarily timelike, a timelike \mathcal{D}_p implies that \mathcal{D}_p points into the future. Hence or within the rest frame in the local rest frame the path $\mathcal{D}_p^2 = 0 = \mathcal{D}_p^2 = 1, 2$.

5. The electromagnetic equation

In the next work we have given a derivation and has derived an expression for the optical character of the medium. So we set up a very general equation in the form of an equation

$$\mathcal{D}_p \mathcal{D}_q \mathcal{A}^p = 0, \quad (5.7)$$

connecting the wavevector \mathcal{D}_p with the coordinate x^p . This generalized equation increases complexity, as far as a generalization is concerned, and optical

number of the orbit is infinite.

The commutator of the action may be interpreted in several ways. If it is defined like σ_2 , we shall be able to write

$$\begin{aligned}\sigma_2 &= \frac{1}{2} (\sigma_1^2 + \sigma_2^2 + \sigma_3^2 + \sigma_4^2 + \sigma_5^2 + \sigma_6^2 + \sigma_7^2 + \sigma_8^2) + \dots \\ \sigma_2 &= \frac{1}{2} (\sigma_1^2 + \sigma_2^2 + \sigma_3^2 + \sigma_4^2 + \sigma_5^2 + \sigma_6^2 + \sigma_7^2 + \sigma_8^2) + \dots\end{aligned}\quad (5.1)$$

It expresses the frequency of phase jumps in terms of their dimension (number, position and time). In the other hand, if it is defined like $(\sigma_2 \sigma_3)^{\frac{1}{2}}$, we shall be able to

$$(\sigma_2 \sigma_3)^{\frac{1}{2}} = \frac{1}{2} (\sigma_1^2 \sigma_2^2 + \sigma_2^2 \sigma_3^2 + \sigma_3^2 \sigma_4^2 + \sigma_4^2 \sigma_5^2 + \sigma_5^2 \sigma_6^2 + \sigma_6^2 \sigma_7^2 + \sigma_7^2 \sigma_8^2) + \dots\quad (5.2)$$

It expresses, by (4.3), the speed of the phase jumps in terms of the direction of propagation, frequency, position and time.

For general mathematical arguments it is best to use the general form (5.1). (5.2) takes the general theory into further nonlinear equations (it is essentially the nonlinear). (5.1) connects the theory with physical ideas, the speed of propagation being expressed in terms of direction of propagation and frequency.

It is a nonlinear commutator (action) function σ_2

$$\sigma_2 = \int \mathcal{L}(\sigma_1, x) = \sigma_2^2 \sigma_3^2 + \dots + \dots\quad (5.3)$$

Now we define the factor σ_2 in terms of the action as a statistical commutator. It is the action, not the action, that is given.

For an σ_2 action of the form

$$\sigma_2 = \int \mathcal{L}(\sigma_1, x) = \sigma_2^2 \sigma_3^2 = \sigma_2^2 \sigma_3^2 + \dots + \dots\quad (5.4)$$

By (4.3) this is equivalent to $\sigma_2 = \sigma_2 \sigma_3$. The relation (5.4) is in general a function of σ_2^2 (the action for dispersion) and a function of $\sigma_2, \sigma_3, \sigma_4$ (the action

for convenience).

Let a topological sector at rank^3 be given

$$\varepsilon \int_{\mathbb{R}^3} \langle \mathbf{D}_\nu, \mathbf{v} \rangle = \sigma_1^2 \lambda_1 \lambda_2 + \sigma_2^2 \lambda_2 \lambda_3 + \sigma_3^2 \lambda_3 \lambda_1 = 0,$$

$$\lambda_1 = \kappa_1^2 \sigma_{II} \sigma_{II}^* + \kappa_2^2 \sigma_{II}^2, \quad \lambda_2 = \kappa_2^2 \sigma_{II}^* \sigma_{II} + \kappa_3^2 \sigma_{II}^2, \quad (5.6)$$

$$\lambda_3 = \kappa_3^2 \sigma_{II} \sigma_{II}^* + \kappa_1^2 \sigma_{II}^2.$$

Let the unit six potential and κ_μ are 1 or given signs in linear potential equations. Because of σ_{II}^2 is also for dispersion. The equation (5.6) reduces to (5.7) if we set $\lambda_1 + \lambda_2 + \lambda_3 = 0$.

4. Relativistic media

Let there be any number of media, each in uniform motion. They may be regarded as systems, as they may exist in one medium. Each medium is, then, treated as its own frame, which is either had possibly (arbitrary) or crystal-like (as media may constitute crystalline media).

The history of media which moves with a velocity of translation and the path of these bodies is, then, as in (5.4) and (5.5), the electro-dynamics equation in the rest frame of the media. The velocity is a simple one. To pass from the local rest frame to a single Minkowski frame from which all the media are treated. This is so

* Cf. Minkowski (1908), p. 200. There is a slight change in notation from Minkowski's ($\sigma^i, \mathbf{I}, \dots$) here the dimension $[\sigma^i, \mathbf{I}]$ stands for σ^i , but the dimension $[\sigma^i]$

by using the homogeneous equations with values to (5.1) and (5.1) in the final two terms. For a reason we have to state as in (5.2)

$$f(\Omega)(\varphi^2, x) = (\sigma_x^2 \varphi_x^2) = 0, \quad (5.1)$$

a homogeneous equation

Let μ_2 be the frequency of a motion with values fixed in relation to the nature of the motion, and the motion is uniform. For an isotropic motion we have

$$f(\Omega)(\varphi^2, x) = (\sigma_x^2 \varphi_x^2) = (\omega^2 - \nu^2) (\sigma_x^2 \mu_2^2) = 0, \quad (5.2)$$

which by putting $\mu_2 = 1$, $\mu_2 = 1, 2$. Here ν is a function of $\sigma_x^2 \mu_2$. It is also the frequency and also of ν_2 . It is also the frequency, subject to

$$\frac{\partial \nu}{\partial \mu_2} \mu_2 = 1. \quad (5.3)$$

For a crystalline motion the principal axes subject to a common motion. From an ideal cell motion, forming an isotropic lattice with $\mu_2 = \nu$ is done first by $\nu_2^{(0)}$, is 1 of the form of the motion we have (using the principal axes as one of the axes)

$$\varphi_1^{(1)} + \varphi_2^{(1)} + \varphi_3^{(1)} = 0, \quad (5.4)$$

and the other components isotropic. Thus, for a general isotropic motion, we have

$$\begin{aligned} f(\Omega)(\varphi^2, x) &= (\sigma_x^2 \varphi_1^{(1)} \varphi_2 \varphi_3) + (\sigma_x^2 \varphi_2^{(1)} \varphi_1 \varphi_3) + (\sigma_x^2 \varphi_3^{(1)} \varphi_1 \varphi_2) \\ &+ \varphi_1^2 + \varphi_2^2 \sigma_x^2 \varphi_3 + (\omega_1^2 - \nu^2) (\sigma_x^2 \mu_2^2) = 0, \\ &+ \varphi_2^2 + \varphi_3^2 \sigma_x^2 \varphi_1 + (\omega_2^2 - \nu^2) (\sigma_x^2 \mu_2^2) = 0, \\ &+ \varphi_3^2 + \varphi_1^2 \sigma_x^2 \varphi_2 + (\omega_3^2 - \nu^2) (\sigma_x^2 \mu_2^2) = 0. \end{aligned} \quad (5.5)$$

Here $\sqrt{g^{(2)}}$ and μ_2 are variables in the output and σ_{μ_2} are functions of $\sigma_{\mu_1} \mu_1$ in case the distribution is not stationary. In case stationarity is independent of σ_{μ_1} , then for the choice of quantities (or a family of models and instead it is to be regarded as a distribution function of σ_{μ_1} .

To follow the propagation of data through cells as considered above we need the amount of input and also the law of reflection and refraction; these will be treated in sections 8 and 10 respectively.

7. Invariant Form of the Control Equations

We have been allowed to set up a frequency-domain equation for the generalized motion of a repulsive motion in case of a solid interface motion, with no change in the wave velocity hypothesis for the case of an unbounded fluid interface.

We shall assume a basic equation between the phase speed $v^{(2)}$ and the frequency $\sqrt{g^{(2)}}$, both measured in the local rest frame. Since we use the usual method for quantizing, this process is an assumption of energy conservation. It is assumed to introduce the definition here, as by the definition

$$v = v^{(2)} / \sqrt{g^{(2)}} \quad (7.1)$$

Then by (4.1)

$$v^2 - 1 = \frac{\sigma_{\mu_2} \sigma_{\mu_1}}{\sigma_{\mu_2} \mu_1^2} \quad \sqrt{g^{(2)}} = -\sigma_{\mu_2} \mu_2 \quad (7.2)$$

By following equation (7.1) we have

$$g \cdot \nabla (\sigma_{\mu_2} \mu_2) = \sigma_{\mu_2} \sigma_{\mu_1}^2 - 2v^2 (1 - \sigma_{\mu_2} \mu_1^2) \sigma_{\mu_2} \mu_2 \quad (7.3)$$

This is the vector equation defining μ_p for a fluid in general motion provided we agree a measure of μ as a function of $\sqrt{p^2}$ and $\rho(x)$ is the proper density of the fluid, so write

$$\mu = \mu(\sigma_p \mu_p, p^2), \quad (7.4)$$

Note that $\rho(x)$ depends on σ_p^2 explicitly and also implicitly through μ ; it is implicit through $\mu_p(x)$ and also through $\rho(x)$, these dependences being ignored from now.

The equation (7.4) has the same form as (5.2) - the difference is that we previously thought of μ_p as constant (vector) fields, and now we think of it as variable (arbitrary) fields.

As a particular case of (7.4) we may take the relativistic formula (cf. S. S. That (7.5) is (5))

$$\mu^{-1} = \sum_{\alpha} \frac{u_{\alpha}^2}{\beta - u_{\alpha}^2} + \sum_{\alpha} \frac{u_{\alpha}^2 v_{\alpha}^2}{v_{\alpha}^2 - \sqrt{1-u_{\alpha}^2}}, \quad (7.5)$$

where u_{α} and v_{α} (the absorption frequencies for the gases at rest) are to be thought as being functions of the density ρ . Now for a radiation source the stress-energy tensor is

$$T_{ik}(0, x) = \sigma_p \sigma_p + (\sigma_p \mu)^2 \sum_{\alpha} \frac{u_{\alpha}^2 v_{\alpha}^2}{v_{\alpha}^2 - (\sigma_p \mu)^2} \quad (7.6)$$

The fields between photons would be an example for absorption. Transmitted or reflected fields produce the same stress-energy

$$\begin{aligned} T_{ik}(0, x) &= \sigma_p \sigma_p + \omega^2 - 2i(\sigma_p \mu)^2, \\ \mu^{-1} &= 4|\sigma_p|^2 + 2|\rho|^2 (\sigma_p \mu)^2. \end{aligned} \quad (7.7)$$

4. Hamilton's partial differential equation. Characteristics as curves.

Let us substitute \mathcal{H}_p for \mathcal{H} , the classical Legendre transform

$$\mathcal{H}_p(q, p) = \mathcal{H} \quad (2.11)$$

as a partial differential equation of the first order for the given function $\mathcal{H}_p(q)$. This is Hamilton's partial differential equation and of third order in it as H.E.D. It is essentially the same as the Hamilton system, in particular, is called the Hamilton-Appell equation.

It was assumed that the problem of solving the H.E.D. subject to initial conditions, the \mathcal{H}_p is solved in a closed form. The actual of solving this problem by means of characteristics is well known, but the argument will be presented here because it is desirable to have it before us in the present solution and it is a type which shows us that the actual system exists.

If $\mathcal{H}_p(q, p)$ is not function of its other variables $q_2 = q_3$, but merely satisfies (2.11), the ordinary differential equation

$$\frac{dq_1}{dt} = \frac{\partial \mathcal{H}_p}{\partial p_1} = \dots = \frac{\partial \mathcal{H}_p}{\partial p_2} = \frac{\partial \mathcal{H}_p}{\partial p_3} = \dots = \frac{\partial \mathcal{H}_p}{\partial p_n} = \mathcal{H}_p \quad (2.12)$$

defines a set of curves in the (q, p) space, the curve passing through any assigned point is that space. These equations may also be written

$$\frac{dq_1}{dt} = \frac{\partial \mathcal{H}_p}{\partial p_1} = \frac{\partial \mathcal{H}_p}{\partial p_2} = \dots = \frac{\partial \mathcal{H}_p}{\partial p_n} = \mathcal{H}_p \quad (2.13)$$

Letting a parameter s label curves along with curves, it is added that (2.12) imply

$$\frac{dq_1}{ds} = 0 \quad (2.14)$$

instead of condition (14) of (17), let us use instead the
 (15) of (16) as a set of conditions that associated with the σ_p and
 from such conditions with the condition σ_p is a distribution of the
 character of the being determined by an initial state and on which σ_p pre-
 sents the function $\int \sigma_p(x)$ has been defined. To get the same as if we divide
 (15) by any function of (15), the only difference being in the parameter λ .

To see that the condition for a variable of (15) is

$$\frac{\partial \sigma_p}{\partial x} = \frac{\partial \sigma_p}{\partial x} \leq 0 \quad (16)$$

(16) is an essential condition to condition also we need to identify the sign with
 the value of the function, since it would be more than a sign of the condition
 (16).

To see that in the case

$$\int \sigma_p(x) = 1, \quad \sigma_p = \sigma_p, \quad (17)$$

and with a solution σ_p taking constant values σ_p on some region Ω .

The solution, as in (15), σ_p would be a state with condition $\sigma_p = 1$

$\sigma_p = \sigma_p = 1$ on (15)

The plot is to choose σ_p on Ω to satisfy

$$\sigma_p \ln \sigma_p = 0, \quad \int \sigma_p(x) = 1, \quad (18)$$

For every displacement in Ω , since there are three degrees of freedom in the dis-
 placement, we have four equations for the condition (18) instead of the number
 of equations is eight. It may happen that we can make σ_p on Ω to be that. To
 investigate this, we can substitute in some cases $\sigma_p = 1$ and find the form of

a Riemannian 1-space in which σ_p are isothetic. The class of (A.7) determined by the spherical projection of σ_p on the sphere is that of $\beta = \text{const}$ [1], and the second states the existence of σ_p as a certain 1-space $\Delta = 0$. As can be seen from (A.7) it, and only it, the spherical projection of $\Delta = 0$ on Π contains the spherical spherical projection of σ_p on Π . And if a solution exists, there can be none other.

Under the operation of the condition of the possible isothericity of (A.7) and of suitable isothericity, we shall assume that a solution exists and that it is unique.

Let the initial values of σ_p on Σ given by (A.7), as provided in the case by (B.1). In general, there may thus be a continuous filling a portion of space-time, but there is no continuous case. We, on Σ , obtain

$$\sum_{\alpha} \sigma_{\alpha} + \dots = 0 \quad (A.8)$$

(we have taken the equation of Σ to be $\beta = 0$), that the sign is not known σ , but it is supposed that σ is less than zero.

We have then a continuous of space extending from Σ . At any point x_p in the domain filled by time as (B.1) we obtain $F(x)$ by

$$F(x) = F(x) + \int_{\Sigma} \sigma_{\alpha} \theta_{\alpha} \quad (A.9)$$

where θ_{α} is the unit vector the way through x_p with $\theta = 0$, and the integral is taken along the way. Now $F(x)$ is the spherical value on Σ , and we are at once that $F(x)$ satisfies the spherical condition.

It is not to be seen that $F(x)$ satisfies the DCE, and to do this we vary x_p (checked).

$$f_{\sigma}^* h_{\sigma} = \sigma(\Delta) + 1 \int_{\sigma}^{\sigma} \sigma_r h_{\sigma} \quad (3.11)$$

the relation being true for the case $\sigma = \text{characteristic value}$. From the case $\Delta(\sigma_r, \Delta) = 2$ in σ_r , it follows that $\Delta(\sigma) = 0$, whence, as before, $\Delta \Delta = 0$.
 Note, in connection with this, the last case of (3.11) is

$$\begin{aligned} \int_{\sigma}^{\sigma} \sigma_r h_{\sigma} &= \sigma(\sigma_r h_{\sigma}) = \int_{\sigma}^{\sigma} \sigma \sigma_r h_{\sigma} = \sigma_r \sigma(\sigma_r) \\ &= \sigma_r h_{\sigma} = \int_{\sigma}^{\sigma} \sigma \sigma_r = \sigma(\sigma_r) \sigma_r \end{aligned} \quad (3.12)$$

$$\sigma_r h_{\sigma} = \sigma(\sigma_r) \sigma_r = \sigma_r h_{\sigma} \quad (3.13)$$

From (3.13) we have that

$$f_{\sigma}^* h_{\sigma} = \sigma_r h_{\sigma} \quad f_{\sigma}^* = \sigma_r \quad (3.14)$$

Let therefore F denote the set of all σ_r such that $\Delta(\sigma_r, \Delta) = 2$.

Now, to obtain a proof of the theorem, we consider an σ_r as that element of the set F which is the image of σ under the mapping $\sigma \rightarrow \sigma_r$ of (3.14). We then write down F as a set of σ_r .

We note that the only element of the set F is σ_r .

$$\int_{\sigma}^{\sigma} \sigma_r h_{\sigma} = \sigma_r \quad \Delta(\sigma_r, \Delta) = 2 \quad (3.15)$$

Let σ_r be the element of the set F which is the image of σ under the mapping $\sigma \rightarrow \sigma_r$. Now to obtain a proof of the theorem, we consider an σ_r as that element of the set F which is the image of σ under the mapping $\sigma \rightarrow \sigma_r$ of (3.14).

9) Extremwerte von $f(x, y, z)$:

Die Werte von f an den NRT-Stationen P_1 und P_2 sind zu ermitteln, so die Werte an den Werten P_3 und P_4 sind zu ermitteln, so:

$$f = f(x, y, z) = 4 \frac{xy + yz}{x^2 + y^2 + z^2} \quad (3.1)$$

• Ist es möglich, dass die Werte $f(x, y, z)$ an den NRT-Stationen P_1 und P_2 sind zu ermitteln, so die Werte an den Werten P_3 und P_4 sind zu ermitteln, so:

$$f(x, y, z) = 4 \frac{xy + yz}{x^2 + y^2 + z^2} = 4 \left(\frac{xy}{x^2 + y^2 + z^2} + \frac{yz}{x^2 + y^2 + z^2} \right) = 4 \quad (3.2)$$

Die Werte f an den NRT-Stationen P_1 und P_2 sind zu ermitteln, so:

$$f(x, y, z) = 4 \frac{xy + yz}{x^2 + y^2 + z^2} = 4 \left(\frac{xy}{x^2 + y^2 + z^2} + \frac{yz}{x^2 + y^2 + z^2} \right) = 4 \quad (3.3)$$

Die Werte f sind:

$$f(x, y, z) = 4 \frac{xy + yz}{x^2 + y^2 + z^2} = 4 \quad (3.4)$$

Die Werte f sind:

$$f(x, y, z) = 4 \frac{xy + yz}{x^2 + y^2 + z^2} = 4 \quad (3.5)$$

Die Werte f sind:

$$f(x, y, z) = 4 \frac{xy + yz}{x^2 + y^2 + z^2} = 4 \quad (3.6)$$

Die Werte f sind: $f(x, y, z) = 4 \frac{xy + yz}{x^2 + y^2 + z^2} = 4$. Die Werte f sind:

normal to the surface, so

$$\mathbf{r} = (0, 0, v_p^2) = v^2 \mathbf{e}_3 \quad (3.7)$$

Thus (3.6) reads

$$\frac{\partial}{\partial t} \mathbf{r}_p = \frac{\partial}{\partial t} (v^2 \mathbf{e}_3) \quad (3.8)$$

The normal to the surface is being carried out with the \mathbf{e}_p frame. Observe that the time \mathbf{r}_p and \mathbf{e}_p (3.8) is an equation connecting \mathbf{v} and $\dot{\mathbf{v}}^2$. Not a strong formula for \mathbf{v} is being used, so

$$\frac{1}{v} = \frac{\partial}{\partial t} \left(\frac{1}{v} \right) \quad (3.9)$$

where v is the phase speed corresponding to frequency ω , and so (3.8) takes us from the geometry of the ray (relating normal to the wave surface) to the wave speed. The wave speed of this Poincaré result is a result of the vector wave equation (3.4) is (3.9).

Although the geometry field is general action is derived by the above equation, it is interesting to mention an example. By (3.1) we have

$$\frac{\partial \mathbf{r}_p}{\partial t} = \mathbf{v}_p - \dot{v}^2 - \dot{v}^2 \mathbf{e}_3 \mathbf{e}_3 \mathbf{e}_3 = \mathbf{v}^2 \mathbf{e}_p + \dot{v}^2 \mathbf{e}_3 \mathbf{e}_3 \mathbf{e}_3 \quad (3.10)$$

$$\frac{\partial \mathbf{r}_p}{\partial t} = \mathbf{v}_p - \dot{v}^2 - \dot{v}^2 \mathbf{e}_3 \mathbf{e}_3 \mathbf{e}_3 = \mathbf{v}^2 \mathbf{e}_p + \dot{v}^2 \mathbf{e}_3 \mathbf{e}_3 \mathbf{e}_3$$

where \mathbf{v}^2 is the direction with respect to $\mathbf{e}_p \mathbf{e}_3$. Substitution in (3.1)

gives the ray velocity. In the limit $\dot{v}^2 \rightarrow 0$, there we have $\mathbf{r}_p = \mathbf{v}^2 + \mathbf{e}_3$, and so in (3.1)

$$\begin{aligned}
 \frac{v^{(2)}}{v} &= 1 + \frac{v_0^{(2)}}{v^2} \frac{v_0^{(1)}}{v} \\
 &= v^2 \frac{v_0^{(2)}}{v^{(2)} v^2 + u u' v^{(2)}} \\
 &= \frac{v_0^{(2)}}{v^2 + u u' v^{(2)}}
 \end{aligned}
 \tag{3.22}$$

to insure that the ray velocity and the group velocity have the same direction (or possibly opposite directions) in the limit, we find, and so have

$$\frac{v_0^{(2)}}{v^{(2)}} = \frac{v^2 + u u' v^{(2)}}{v} + u = v^{(2)} \frac{v_0}{v^{(2)}} = \frac{v_0}{v^{(2)}} (1 + v^{(2)})
 \tag{3.23}$$

The right hand side is a well known form for $v_0^{(2)}$, equivalent to (3.5), and thus as verified in this particular case the more general result established above.

This brief discussion of group velocity has been included here in order to give the ideas of the present paper to those workers in physics. Not having seen this discussion, one may very well think about group velocity and think in terms of ray velocity, to which it appears to be equivalent. In the theory of optics as ray group velocity seems an artificial addition. Variation of the frequency ω of a wave in ordinary geometrical optics. Due to the present theory, variation of frequency is allowed, even in the statement principle (3.14) for rays, referring to one of the variables to be varied. Thus the essential idea of group velocity is built into the present theory.

The inequality (3.6) is the condition that the group velocity should not exceed c .

3. Noting

We use quantum photon gas statistical system by the following statement:
 Each one is a possible state of a system, and the probability density of the system is

$$P_j = \frac{1}{Z} e^{-\beta \epsilon_j} \quad (15.1)$$

where ϵ_j is the energy of a system and $\beta = \frac{1}{k_B T}$ is the inverse of the temperature multiplied by Boltzmann's constant.

This system is given to the photon gas in practical system because the role of the particle is quantum statistics. It is a boson gas subject with constant energy. It is a scalar photon when does not consider polarization, but in the context of theoretical calculation does not matter both.

The temperature T_j in (15.1) has the dimension of energy. We say that

$$T_j = \frac{\epsilon_j}{k_B} \quad T_j = 1 \text{ eV} \quad (15.2)$$

where T_j is the temperature of the state and ϵ_j the energy. Now by (15.1) and (15.2) we have

$$P_j = \frac{1}{Z} e^{-\frac{\epsilon_j}{k_B T}} = e^{-\frac{\epsilon_j}{k_B T}} \quad Z = \sum_j e^{-\frac{\epsilon_j}{k_B T}} = e^{-\frac{\epsilon_j}{k_B T}} \quad (15.3)$$

where ϵ_j is the energy density

The stationary condition (15.1), addition to a rate, may be written

$$\int \epsilon_j^2 \rho_j \omega_j = 0 \quad \int \epsilon_j \rho_j \omega_j = 0 \quad (15.4)$$

The integral involving here may be written (it is calculated by using the eqn.)

$$\int \epsilon_j^2 \rho_j \omega_j = 0 \quad \int \epsilon_j \rho_j \omega_j = 0 \quad (15.5)$$

so that we remember it as an action integral. The integral is along any world line (not necessarily a ray) we like:

$$S = \int_{x_1}^{x_2} \mathcal{L} dx = \int_{x_1}^{x_2} (\mathcal{L}_0 + \mathcal{L}_1) dx \quad (12.4)$$

where $\mathcal{L}(x)$ is the classical Lagrangian. If we would like to derive from a world of a certain phase S we have to ask about the real phase S and with the same phase then by (12.4) we have $\mathcal{L}(x) = \dot{x} \mathcal{L}_0 + \mathcal{L}_1$ and therefore

$$\mathcal{L}_0 = \dot{x} \mathcal{L}_1 \quad (12.5)$$

to write. On action between adjacent phase surfaces of the same phase is equal to Planck's constant. (12) The process of production of particles, from \mathcal{L}_0 to \mathcal{L}_1 .

We can refer the question: "Is a photon a proper mass?" and a question is a \mathcal{L}_0 particle production without a definition of proper mass. It is natural, perhaps, to take the definition appropriate in a relativistic particle, for which the proper mass m_0 is given by the ratio of energy E and momentum p_0 by

$$m_0^2 c^4 = E^2 - p_0^2 c^2 \quad (12.6)$$

Then we substitute in the right hand (12.6) we get

$$m_0^2 c^4 = \mathcal{L}_0^2 \mathcal{L}_1^2 c^2 - \mathcal{L}_1^2 / \mathcal{L}_0^2 = -\mathcal{L}_1^2 \mathcal{L}_0^{-2} \quad (12.7)$$

For a massless particle $m_0 = 0$ [cf. (12.6)]. So in the arbitrary case of a relativistic particle, for which $m_0^2 > 0$ is equivalent \mathcal{L}_0^2 is positive, it means \mathcal{L}_1^2 must be and is negative.

(12) We can be useful to speak of the proper mass of a photon at all. We

should be kept for a definition which makes $\alpha \neq 0$. For a moment let $\alpha = 0$ and for a section in which $\alpha \neq 0$, we have another set of things to do apart from our definition, as that

$$\alpha = \frac{\Delta \alpha}{\alpha^2} \left(\frac{\alpha^2}{\alpha^2} - 1 \right) \beta + \frac{1}{\alpha} \langle \alpha_p \alpha_p \rangle \beta. \quad (11.10)$$

11. Definition of a physical system. DEFINITION:

Let us try to find a picture of the interaction of a photon by an atom in terms of geometrical optics.

First we recall the content of Section 6. Given a junction Σ with β enclosed in Ω , we sought to find β and β' that we had to solve (9.7), that

$$\nabla_{\alpha} \langle \alpha_p \rangle = \alpha \beta, \quad \int_{\Sigma} \langle \alpha_p \rangle \cdot \alpha = \beta, \quad (11.11)$$

where $\beta = \beta'$ on Σ . If these equations can be solved and the straightened line (5.4) is normal, we get rays emanating from Σ and filling a 3-dimensional region of space-time.

Suppose now that Σ is a junction located in a junction. Then the conditions (11.11) are affected in three respects (instead of two), but as (11.11) is not more than three equations) we get ω^3 values of $\langle \alpha_p \rangle$ at each event on Σ . That we have ω^3 rays emanating, enough to fill a 3-dimensional region of space-time as before.

If Σ is a world line (1-space), then we have only two equations in (11.11) and hence ω^2 values of $\langle \alpha_p \rangle$ at each event on Σ . There are ω^2 rays

an Ω , and we again we get rays adequate to fill a 4-dimensional region of space time. If this world line Ω is regarded as the world line of an emitting atom (the frequency of emission defining Ω as Ω), we can appreciate better why we do not think of a photon as both of these rays, but only one photon for the whole process. The full amplitude over possible positions of this one photon, and the associated values of $\nabla_{\mu} \psi$, its possible momentum-energy derivatives. We can use to this process a solution of Sakurai's uncertainty principle. The initial frequency is assigned and the initial time completely undetermined, and the other solution, momentum and energy of the photon both that degree of indeterminacy restricted to the fact that, although we have a definite picture of rays in space-time, we do not know where one the photon has chosen to interact until it shows its presence by colliding with matter.

If ν_0 is the frequency of the source or atom and ν the observed frequency of radiation, then we have as in (11.1) the two functions

$$\psi_0 = \nu_0 e^{-i\omega_0 t} \quad \psi = \nu e^{-i\omega t} \quad (11.2)$$

to be satisfied by ψ_0 on the world line of the source, and the details of the ψ to be found by solving (8.4) subject to these initial conditions. The phase function θ is then given by (3.9) with $\theta(x) = -i\omega t / \hbar = \hbar \nu_0 t$.

Interference phenomena can be treated only strictly in geometrical optics.

If, by means of the slit-slit or interferer, we allow two paths to the rays, we can distinguish (if the slits are the two rays differ by δx and distance if they differ by $(\delta + \delta) x$, δ being an integer, then the waves corresponding to differences of phase values of amount $2\pi\delta$ and $2\pi\delta + \pi$, respectively,

In terms of probability, we may say that, in the 710-0 case, it is very probable that the photon will induce one or other of the two jets and is the parent lepton that, in the 710-1 case, it will induce neither.

Success in production of a photon will not result in the electron. This is explained in appendix A. The possible values of \mathcal{D}_p are subject only to $\mathcal{D}_p(\mathcal{D}^2, x) = 0$. (E. 44). We have (III.1) with the first equation deleted. This means that every jet is a 1-jet (total value of \mathcal{D}_p and its total angular momentum) and we have, in the same circumstances, a set of jets forming a 4-dimensional vector of spin-0's. Let us consider the question, treated in technical notation with $\mathcal{D}_p(\mathcal{D}^2, x)$ as in (I.1).

Writing $\mathcal{D}_p = \mathcal{D}_p / \mathcal{D}^2$, $q = \mathcal{D}^2 - 1$, $\mathcal{D}^2 = \mathcal{D}_p^2 + 4\mathcal{D}_p \mathcal{D}_p^2 + 1$, we have

$$\mathcal{D}_p = \frac{\mathcal{D}_p^2}{\mathcal{D}_p^2} = \mathcal{D}_p - 4\mathcal{D}_p \mathcal{D}_p^2 \mathcal{D}_p - 8\mathcal{D}_p^2 \mathcal{D}_p^2 (\mathcal{D}_p^2 \mathcal{D}_p^2)^2 + \dots \quad (III.1)$$

Let us use the total rest frame, so that $\mathcal{D}_p^2 = 0$.

$$q \mathcal{D}_p(\mathcal{D}^2, x) = \mathcal{D}_p^2 \mathcal{D}_p^2 + (q - 4) \mathcal{D}_p^2 = 0, \quad (III.2)$$

and (III.1) becomes

$$\mathcal{D}_p = \mathcal{D}_p^2 + \mathcal{D}_p^2 + (q - 4) \mathcal{D}_p^2 + 8 \mathcal{D}_p^2 \mathcal{D}_p^2, \quad (III.3)$$

and (III.4) gives

$$\mathcal{D}_p^2 \mathcal{D}_p^2 = (q - 4) \mathcal{D}_p^2 = 0, \quad (III.4)$$

whereas \mathcal{D}_p^2 is to be regarded as a function of \mathcal{D}_p^2 determined by the amount of (III.3), so that \mathcal{D}_p^2 varies randomly and also independently to q and \mathcal{D}_p^2 .

It appears in the next part a 3-dimensional region of operations, obtained by joining the points of \mathcal{D}_p^2 space to the points in the space of (III.4). Let us take

the vector is non-dissipative, then $\psi = \psi_1 + \psi_2$ being a function of ρ only and hence a function of position only. The second of (11.10) reduces to

$$\dot{\psi}_1 + (1 + \alpha) \psi_2 = 0 \quad (11.17)$$

and (11.15) reads

$$\dot{\psi}_1 \psi_{1\rho} + \alpha^2 \dot{\psi}_2^2 = 0 \quad (11.18)$$

Now if α were constant within the null cone (and it should be to meet the speed of the photon limit $\alpha \leq 1$ if $\beta \leq 1$). The fact that α is a non-zero function of ρ may then be thought of as a non-dissipative process within (11.10) in a three-dimensional region in spacetime. This is of course true in particular for a reflector, the case being that the null cone $(\dot{\psi}_1 \psi_{1\rho} = 0)$

14. The case of reflection and refraction

Let Σ be the locus of a moving surface separating media \mathcal{M}^+ & \mathcal{M}^- . With light passing from \mathcal{M}^+ into \mathcal{M}^- (refraction) or is reflected back into \mathcal{M}^+ , a discontinuity occurs, and the variational equation (11.12) leads at once to the equation

$$\delta \mathcal{O}_\rho^+ = \delta \mathcal{O}_\rho^- + \delta h_\rho = 0 \quad (11.19)$$

for all displacements h_ρ in Σ . \mathcal{O}_ρ^+ and \mathcal{O}_ρ^- being the corresponding densities in \mathcal{M}^+ and \mathcal{M}^- respectively of the mass of reflection or refraction, respectively.

$$\mathcal{Q}_p^* = \mathcal{Q}_p^* + k \mathcal{K}_p, \quad (11c)$$

where \mathcal{Q}_p is the unit vector in \mathcal{R} pointing from \mathcal{W} into \mathcal{D}_p , and k is an interaction factor. This is the law of reflection and refraction. It tells us that the incident in the domain-frequency domain is equal to the history of the system of equations.

Further, \mathcal{Q}_p^* is given, not only the unit vector, but also the reflection or refraction vector, which is (11c) for incident in the time domain. $\mathcal{Q}_p^* + k \mathcal{K}_p$ by definition is a unit vector is equal to the average of a unit vector of \mathcal{D}_p .

$$\int_{\mathcal{D}_p} (\mathcal{Q}_p^* + k \mathcal{K}_p) d\mathcal{V} = 0, \quad (11d)$$

with the reflection vector to be the domain-frequency domain for \mathcal{W} , and

$$\int_{\mathcal{D}_p} |\mathcal{Q}_p^* + k \mathcal{K}_p| d\mathcal{V} = 1, \quad (11e)$$

with unit be satisfied not only by \mathcal{Q}_p^* , but also by $\mathcal{Q}_p^* + k \mathcal{K}_p$.

The boundary reflection, which is a vector (11c) and (11d) process unknown. If long time, we have still in averaged state the reflected unit vector same like \mathcal{W} . The boundary reflection, which is the same, with (11c) and (11d).

In view of this, we try to simplify the state, as far as possible by choice of the state of reference. Suppose, in general, that the speed of the unit vector of incident is function ϕ , the vector \mathcal{K}_p is constant. Let us take the \mathcal{K}_p with its direction, so that $\mathcal{K}_p = 1 = \mathcal{K}_1 = \mathcal{K}_2 = \mathcal{K}_3 = 1$. Then (11c) can

$$\mathcal{Q}_1^* = \mathcal{Q}_1^* + k, \quad \mathcal{Q}_2^* = \mathcal{Q}_2^*, \quad \mathcal{Q}_3^* = \mathcal{Q}_3^*, \quad \mathcal{Q}_4^* = \mathcal{Q}_4^*, \quad (11f)$$



Let us start by our incident wave at fixed position. The wave is not stationary. Supposing the incident wave ψ_0 (we use as a fixed wave in (1), (2), (3) and (4)) may be written

$$\begin{aligned} \psi_0 &= \psi_0^+ + \psi_0^- + \psi_0^0 + \psi_0^1 + \psi_0^2 + \dots \\ \psi_0^+ &= \psi_0^+ e^{i(k_0 x - \omega_0 t)} + \psi_0^+ e^{-i(k_0 x - \omega_0 t)} + \dots \end{aligned} \quad (15.8)$$

These equations are to be solved for ψ , the field in the case of reflection and the wave in the case of refraction.

Let us suppose the wave comes from the left and incident, so that by (15.8)

$$\begin{aligned} \psi_0 &= \psi_0^+ e^{i(k_0 x - \omega_0 t)} + \psi_0^- e^{-i(k_0 x - \omega_0 t)} + \dots \\ \psi_0^+ &= \psi_0^+ e^{i(k_0 x - \omega_0 t)} + \psi_0^+ e^{-i(k_0 x - \omega_0 t)} + \dots \end{aligned} \quad (15.9)$$

where $\omega^2 = c^2 k^2$ are the refractive indices of the media and $\mu_1^2 = \mu_2^2 = \mu_0^2$. With this condition we have a relation (15.9) and therefore

$$\mu_1^2 = \mu_2^2 = \mu_0^2 \quad (15.10)$$

The boundary conditions at $\mu_1^2 = \mu_2^2$ from the Maxwell equations in a homogeneous medium are satisfied, and we may consider the quantization of the field with an arbitrary wave ψ_0 (15.8) by assuming that

$$\mu_1^2 = \mu_2^2 = \mu_0^2, \quad \mu_1^2 = \mu_2^2, \quad \mu_1^2 = \mu_2^2 \quad (15.11)$$

This may be called the quantum theory of reflection for a pair of media (Fig. 6). It is for the media where the wave is in the plane $\mu_1 = \mu_2 = \mu_0$ with total and partial reflection. Denoting the wave number by ω , we have

$$\begin{aligned} \mu_1^2 = \mu_2^2 = \mu_0^2 = \mu_0^2 = \mu_0^2, \quad \mu_1^2 = \mu_2^2 = \mu_0^2 = \mu_0^2 \\ \mu_1^2 = \mu_2^2 = \mu_0^2, \quad \mu_1^2 = \mu_2^2 = \mu_0^2 \end{aligned} \quad (15.12)$$

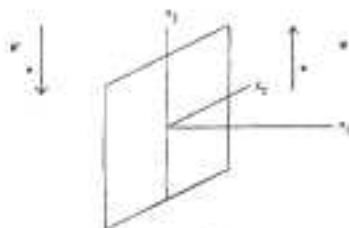
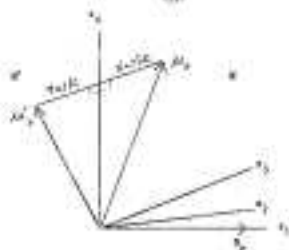


Fig. 6

Standard form of equations for a pair of nuclei.

Upper diagonal equation.

Lower diagonal equation.

and (16.7) become

$$2\Omega = \gamma_4 \Omega_4^2 - (k^2 - 1) \gamma^2 \Omega_1^2 \omega^2 + 4\Omega_4^2 \gamma^2 = 0, \quad (16.11)$$

$$2\Omega = \gamma_2^2 \Omega_2^2 - (k^2 - 1) \gamma^2 \Omega_3^2 \omega^2 + 4\Omega_3^2 \gamma^2 = 0.$$

Invoking the equations (16.4), to be satisfied respectively for reflection and transmission, we

$$\Omega_1^2 + k^2 - \Omega_2^2 + \Omega_3^2 - \Omega_4^2 + (k^2 - 1) \gamma^2 \Omega_1^2 \omega^2 - 4\Omega_3^2 \gamma^2 = 0, \quad (16.12)$$

$$\Omega_1^2 + k^2 + \Omega_2^2 + \Omega_3^2 + \Omega_4^2 - (k^2 - 1) \gamma^2 \Omega_1^2 \omega^2 + 4\Omega_3^2 \gamma^2 = 0.$$

Similar relations, for the modes of $k < 1$ is necessary and sufficient for

$$k^2 - 1) \gamma^2 \Omega_1^2 \omega^2 - 4\Omega_3^2 \gamma^2 - \Omega_2^2 - \Omega_3^2 - \Omega_4^2 \geq 0, \quad (16.13)$$

or equivalently by the sum of (16.11)

$$\Omega_1^2 + \gamma^2 [(k^2 - 1) \Omega_1^2 \omega^2 - 4\Omega_3^2 \gamma^2] - (k^2 - 1) \Omega_1^2 \omega^2 + 4\Omega_3^2 \gamma^2 \geq 0, \quad (16.14)$$

or in terms of their relation by (16.7)

$$\frac{d^2 \gamma^2}{d\omega^2} = \gamma^2 [(k^2 - 1) \frac{\gamma^2}{\omega^2} + 1] - (k^2 - 1) \frac{\gamma^2}{\omega^2} - 1 \geq 0 \quad (16.15)$$

the function γ^2 decreases as a wave factor (γ^2), although still remains positive in ω^2 (Appendix).

Since the wave function is arbitrary, reflection from ω^2 (for $k > 1$) can take place. But there is a failure condition for reflection. It may be possible

the condition to pass from \mathcal{W}^+ into \mathcal{W}^- , and the condition for this is $\eta_1 > 0$ where, by (5.1),

$$\eta_1 = 1 + \frac{W_1^2 \sin^2 \theta}{W_2^2 - W_1^2} \quad (11.18)$$

This we are to calculate by differentiating the expression in the first line of (11.12). Recalling \mathcal{W} non-degeneracy for statistics, we get

$$\begin{aligned} \eta_1 &= 1 + \frac{\sigma_1^2}{\sigma_2^2 + (1-\alpha)^2 - \alpha^2} \frac{W_1^2 (W_2^2 \sin^2 \theta - 1 \sigma_2^2)}{W_2^2} \\ &= 1 + \frac{\sigma_1^2 + \alpha}{\sigma_2^2 + (1-\alpha)^2 - \alpha^2} \frac{W_1^2 (W_2^2 \sin^2 \theta - 1 \sigma_2^2)}{W_2^2} \\ &= \frac{2}{\alpha} \frac{W_1^2 + \alpha}{(1-\alpha)^2 - \alpha^2 + \frac{2W_1^2}{\alpha} \frac{\alpha}{2} + 1} \quad (11.19) \end{aligned}$$

Now assuming (11.11) satisfied, the first of (11.23) gives the real values of $(W_2^2 - \alpha)$, and the solution of the other. The value η_1 positive and the other value is negative. This means that the satisfaction of (11.17) (or equivalently (11.22) or (11.23)) means the existence of a reflected ray proceeding from \mathcal{W}^+ into \mathcal{W}^- , and only one such ray. If dispersion is omitted, there is no such double result in general.

Actually, the case of ray (11) in (11.13) would be that $\alpha < 0$, so that we have $\eta_1 < 0$. We see now that the condition for total reflection is

$$\frac{\alpha^2 \sin^2 \theta}{\alpha^2} = \alpha^2 \left[(1-\alpha)^2 - \alpha^2 + \frac{W_1^2 \alpha}{\alpha^2} + \alpha^2 \right] = (1-\alpha)^2 + \left[\frac{W_1^2 \alpha}{\alpha^2} - \alpha^2 \right] < 0 \quad (11.20)$$

It is recalled that v_1^i is the component of incident phase velocity normal to the surface of separation (fixed in our standard frame) and v_2^i is the component in the direction of motion of the fluid medium U . v is the speed of either medium.

We now insert (11.10) by reducing both media to rest, so that $u = 0$, $\beta = 1$, $v^i = v_1^i$, and $U = v_2^i / v$, where α is the angle of incidence. Then (11.10) becomes

$$v^2 \exp^2 \alpha + v^2 - v^2 \leq 0, \quad (11.10')$$

which is immediately recognized as the discriminant formula for total reflection.

To discuss reflection at non the surface of (11.11) instead of the first, we, for simplicity make the standard frame of reference, the second of (11.12) instead of the first. All σ_j^i satisfy the second of (11.11), and as the equation simplifies to

$$|\sigma_j^i|^2 + v^2 - v^2 \leq 0, \quad (11.10'')$$

this gives the set relations

$$v = 0, \quad v = -2|\sigma_j^i|. \quad (11.11')$$

The first is ruled out (11.10'') gives the correct sign; thus the second is OK for the reflected stress-frequency tensor σ_j^i .

$$\sigma_1^1 = -\sigma_1^1, \quad \sigma_2^2 = \sigma_2^2, \quad \sigma_3^3 = \sigma_3^3, \quad \sigma_4^4 = \sigma_4^4 - \quad (11.12')$$

This is the very simple law of reflection with the standard frame of reference in rest.

Thus the momentum-energy 4-vector of a photon is $p_\mu = h \sigma_\mu^0$, and this is

cannot directly by reflection or refraction, we conclude that a δ -impulse acts on the piston and (in case of the conservation of momentum and energy) that an equal and opposite δ -impulse acts on the mirror or piston. By (11.1) the δ -impulse on the piston is

$$\mathcal{I}_p = + \delta \delta Q_p, \quad (11.13)$$

it being understood by (11.13), (11.1) for reflection and by (11.13), (11.1) for refraction. For a front mirror this gives an impulse of magnitude $-\delta \delta Q_p$ with its front component. For a pair of acoustic media referred to their standard frame the energy of a piston is conserved by reflection or refraction, and the only component of momentum is its normal to the normal to the (front) surface of separation.

15. Derivation of the Hamiltonian equations

The theory of this paper has been developed by methods very close to those used by Newton in the geometrical optics, not if I did not associate the results given by the wave with familiar concepts of Hamiltonian optics the reader might feel we are that well to have developed in it first the Hamiltonian equations of a particle. This is not a substitution of Hamiltonian optics [with the replacement of ordinary time by proper time]. It is applying Hamiltonian dynamics of the form familiar to Newtonian theory.

We had not to add the radiation in by means of a hierarchy of solutions, we being considering systems used in the present paper and similar those entirely contained in *SPACETIME*, with the Liouville theorem. The ANALYSIS of

imposition to get from (10) for most $\alpha = 0, 1, 2$,

	Quantum comm.	Classical Poisson	Canonical Poisson
Coordinate	x_p	$\{x_p, x_p\} = 0$	x_p
Momentum	p_q	$\{p_q, p_q\} = 0$	p_q
Coordinate	x_q	$\{x_q, x_q\} = 0$	x_q
Momentum	p_p	$\{p_p, p_p\} = 0$	p_p
Mixed	$\{x_p, p_q\} = \delta_{pq}$	$\{x_p, p_q\} = \delta_{pq}$	$\{x_p, p_q\} = \delta_{pq}$

Equation of motion (11.5)

$$\dot{x}_p = \frac{\partial H}{\partial p_p} = \frac{\partial}{\partial p_p} \left(\frac{1}{2} p_p^2 + \dots \right) = p_p \quad (11.11)$$

$$\dot{p}_q = -\frac{\partial H}{\partial x_q} = -\frac{\partial}{\partial x_q} \left(\frac{1}{2} p_q^2 + \dots \right) = -p_q \quad (11.12)$$

Hamilton's principle

$$\delta \int \mathcal{L} dt = 0, \quad \delta \int (p_p \dot{x}_p - H) dt = 0 \quad (11.13)$$

Note that in (11.11) and (11.12) the familiar Hamiltonian equations of motion and action principle are the weak formulations of the corresponding formulas in the present formalism.

In standard quantum theory (Sternberg) we are used to starting with a Lagrangian $\mathcal{L} = \mathcal{L}(q, \dot{q}, t)$ where $q^i = dq^i/dt$ and Hamilton's principle

* We refer to \mathcal{L}_2 in (11.7)

$$\int_{\mathbb{R}^n} \psi(x) dx = 0, \quad (15.5)$$

We apply the asymptotic formula by

$$\psi_p = \mathcal{R}(\psi_p) + \mathcal{O}_p(\epsilon^2), \quad (15.6)$$

and write Cauchy's formula for $\mathcal{R}(\psi_p)$ as before, the Riemannian by

$$\mathcal{R}(\psi_p(x)) = \int_{\mathbb{R}^n} \psi_p(x') dx' = \mathcal{R}(\psi(x' + \epsilon^2 x)). \quad (15.7)$$

Then

$$\mathcal{L}(\psi_p) = \int_{\mathbb{R}^n} \psi_p(x) dx = \mathcal{L}(\psi), \quad (15.8)$$

and we calculate the asymptotics of Laplace's integral in the form (15.7) and the action principle written down in (15.2).

In the other hand we apply also (15.6) to $\psi_p(x, \epsilon^2 x) = \mathcal{O}_p(\epsilon^2)$ by

$$\psi_p = \mathcal{R}(\psi_p) + \mathcal{O}_p(\epsilon^2), \quad (15.9)$$

write the $\mathcal{R}(\psi_p)$ and define \mathcal{L} by

$$\mathcal{L}(\psi(x' + \epsilon^2 x)) = \int_{\mathbb{R}^n} \psi_p(x') dx' = \mathcal{L}(\psi(x, \epsilon^2 x)). \quad (15.10)$$

Since to apply (15.8) requires a little differentia.

This second idea is not the usual one in classical mechanics, but it is essentially the idea we used earlier here: it is able to obtain a formula for the action principle. In view of the difficulty, we can be free by the action of (15.7), (15.8), but it is better (having seen the connection with stochastic quantization) to go back to our original method, which takes $\mathcal{L}(\psi(x, \epsilon^2 x))$ in a general form and not in the special form $\mathcal{L}(\psi) = \mathcal{L}(\psi)$.

The knowledge is contained in the equations (ref. [24]), p. 64. I have written ϵ instead of ϵ^2 .

14. Substitution of the vector function $\mathcal{G}(\mathcal{M}, \mathcal{P})$ for an arbitrary vector in general action:

For an arbitrary vector in general action we have the corresponding equation (15.1)

$$\mathcal{G}(\mathcal{M}, \mathcal{P}) + \mathcal{G}(\mathcal{M}) + \mathcal{G}(\mathcal{M}, \mathcal{P}^2) = 0, \quad (16.1)$$

where the first member of the last expression is zero

$$\mathcal{L}_\mu \mathcal{P}_\mu = 0, \quad (16.2)$$

and where

$$\mathcal{G}^2 = 0 = \mathcal{G} = \mathcal{G}[\mathcal{G}(\mathcal{M}), \mathcal{P}]. \quad (16.3)$$

We shall denote by \mathcal{G} the partial derivative of \mathcal{G} with respect to $\mathcal{G}(\mathcal{M})$,

We shall find \mathcal{G} is the integrative vector, applied in the ordinary way for the least root theory, and that is that from the formulae $\mathcal{G}^2 = -\mathcal{G}(\mathcal{M})$.

If we are not given the explicit form of the components of \mathcal{G} in $\mathcal{G}(\mathcal{M})$,

we cannot have to carry out explicitly the determination of the action function $\mathcal{I}(\mathcal{M}, \mathcal{P})$ as mentioned in the preceding section. Now if we are given that

from the relation we may not suppose. We may say at least analyze the

problem with a view to applying calculations in the next section for a method which is only slightly approximate. Then we proceed without approximation.

The basic equations are (13.2) with (11.10). We have thus

$$\mathcal{G} \mathcal{L}_\mu = \frac{\partial \mathcal{I}}{\partial \mathcal{P}_\mu} = \mathcal{P}_\mu - \mathcal{P}_\mu \mathcal{G}(\mathcal{M}) - \mathcal{G} \mathcal{P}_\mu \mathcal{G}(\mathcal{M}^2), \quad (16.4)$$

to be solved for \mathcal{G} . Multiplication by \mathcal{P}_μ shows, since $\mathcal{P}_\mu \mathcal{P}_\mu = -1$,

$$\mathcal{G}(\mathcal{L}_\mu \mathcal{P}_\mu) = -\mathcal{G}(\mathcal{M})(1 + \mathcal{G}) = \mathcal{G} \mathcal{G}(\mathcal{M}^2), \quad (16.5)$$

$$\theta \alpha_r = \frac{\partial \Omega}{\partial \sigma_r}, \quad \Omega(\sigma, x) =$$

with

$$\alpha_r \alpha_r = -1.$$

The plan is to solve the first of (13.8) (four $(\theta \alpha_r, x_r)$), substitute these values in the second, and express Ω as a function of (α_r, x_r) ; hence we obtain Ω as a function of (α_r, x_r) . Substitution in the last of (13.9) gives

$$f = f(x, \alpha).$$

Take any timelike world line with unit tangent

$$f(x, \alpha) ds = -e^{-1} \sigma_r \frac{\partial \Omega}{\partial \sigma_r} ds =$$

and so we reconcile the action principle $\delta \int$

or, by (14.11),

$$\begin{aligned} \dot{\theta} &= \dot{\theta}(x, \dot{x}) = \frac{1}{2} \dot{\theta}^2 \dot{x}^2 (2\dot{x}\dot{\mu})^2 \\ &= \frac{1}{2} \dot{\theta}^2 \dot{x}^2 \dot{\mu}^2. \end{aligned} \quad (14.16)$$

It is easy to see further that (14.16) is a function of \dot{x} and $\dot{\mu}$, hence that the functions $\dot{\theta}$ and $\dot{\mu}$ are inverse.

The one key equation in the above work are (14.5), to be solved for $\dot{\theta}(\dot{\mu})$, and (14.11), to be solved for $\dot{\theta}$. It is this that we are liberty to use partial differentiation to any required calculation.

We can hence concentrate our algebraic attention to a single equation by elimination of $\dot{\theta}$ between (14.5) and (14.11). By (14.5) we have

$$\dot{\theta}(\dot{\mu})\dot{\mu} = \frac{1}{2} \dot{\theta}^2 \dot{\mu}^2 + (2 + \dot{\theta}^2) \dot{\mu}. \quad (14.17)$$

Substituting in (14.11) and dividing by $\dot{\theta}^2 \dot{\mu}^2$ we get

$$\frac{\frac{d\dot{\theta}(\dot{\mu})}{d\dot{\mu}} \dot{\mu} - \dot{\theta}(\dot{\mu})}{\dot{\theta}(\dot{\mu})^2} (2 + \dot{\theta}^2) \dot{\mu} + 1 + \frac{\dot{\mu}^2}{\dot{\theta}^2} - (2 + \dot{\theta}^2) = 0. \quad (14.18)$$

This equation has to be solved for $\dot{\theta}$ (see however that $\dot{\mu}$ and $\dot{\theta}^2$ are some functions of $\dot{\theta}$ and $\dot{\mu}$). When $\dot{\theta}$ has been found, $\dot{\mu}$ is given by (14.15). By (14.16) the Lagrangian or action function is

$$\dot{\theta}(x, \dot{x}) = (2\dot{\mu})^2 \frac{\frac{1}{2} \dot{\theta}^2 \dot{\mu}^2}{\frac{1}{2} \dot{\theta}^2 \dot{\mu}^2 + 1 + \dot{\mu}^2}. \quad (14.19)$$

The case of a uniformly moving is discussed. For such a vector $\dot{\theta}^2 = 1$ and $\dot{\theta}$ disappears from (14.11), which reduces to

$$\dot{\mu}(\dot{\mu})\dot{\mu}^2 = 2 + \dot{\mu}^2. \quad (14.20)$$

It might be a function of $\dot{\mu}$ only. In fact, the correct guess here for a

This is to be regarded as an equation for $(\sigma\mu)$, which appears not only explicitly, but also in q and q' . Let the solution be

$$(\sigma\mu) = \theta [\theta(\alpha\mu), \rho], \quad (14.6)$$

a function of two variables, now supposed known. Then by (14.4)

$$\sigma_r^- = \theta \alpha_r + \mu_r (q\theta + \frac{1}{2}q' \theta^2); \quad (14.7)$$

but by (14.5)

$$q\theta + \frac{1}{2}q' \theta^2 = \theta(\alpha\mu) - \theta, \quad (14.8)$$

and so

$$\sigma_r^- = \theta \alpha_r + \mu_r [\theta(\alpha\mu) - \theta]. \quad (14.9)$$

Hence

$$\begin{aligned} (\sigma\sigma) &= -\theta^2 + 2\theta(\alpha\mu)[\theta(\alpha\mu) - \theta] - [\theta(\alpha\mu) - \theta]^2 \\ &= \theta^2 [(\alpha\mu)^2 - 1] - \theta^2. \end{aligned} \quad (14.10)$$

Our instructions are to substitute for σ_r^- in (14.1), which is the second of (13.9), and we do this by substituting (14.6) and (14.10); we get

$$\theta^2 [(\alpha\mu)^2 - 1] - (1+q)\theta^2 = 0. \quad (14.11)$$

Now q is a function of θ and ρ , and θ is a function of θ , $(\alpha\mu)$ and ρ ; thus (14.11) is an equation to determine θ as a function of $(\alpha\mu)$ and ρ ; let the solution be

$$\theta = K [(\alpha\mu), \rho]. \quad (14.12)$$

By the last of (13.9) combined with (14.4) we have

$$\theta f = -\sigma_r^- \frac{\partial \Omega}{\partial \sigma_r^-} = - [(\sigma\sigma) - q(\sigma\mu)^2 - \frac{1}{2}q'(\sigma\mu)^3], \quad (14.13)$$

15. Matrix Function (Determinant) for a special class of matrix

In the following theorem (2.14) let \sqrt{b} be some, possibly not rational, only the linear term, so $\rho = [\rho, 0, 0]$.

$$x^2 - 1 = x + \rho + B(x) e^{1/x} \quad (2.12)$$

then

$$B(x) = \sum_{k=1}^{\infty} b_k x^{-k}, \quad B(\rho) = \sum_{k=1}^{\infty} b_k \rho^{-k} \quad (2.13)$$

Here $B(x)$ is small. $B = 0$ gives a non-dispersive medium, and as we know of the formula (15.1) or inferred is a slightly dispersive medium.

Let us apply the method of Section 14 to find the matrix function (determinant) of a matrix for above (2.11) form. Our work applies in particular to a slightly dispersive medium (B small), but the generalizations are better, and it is clearer to accept (15.1) as an exact formula and solve the general problem based on the existence of B . Accordingly in the notation of Section 14 we have

$$a_1 + a_2 + B(x) = 0, \quad a_1 = 2(\sigma^2 \rho)^2, \quad a_2 = -2(\sigma^2 \rho) \quad (2.14)$$

So we apply to (2.12), and for the value (15.1) it is suitable in B . Accordingly the matrix can be derived and explicitly.

To simplify the writing, we note that $(\sigma^2 \rho)$ is an invariant, and the value can be obtained by using the total root Dues - we have

$$(\sigma^2 \rho) = \frac{1}{2} \omega_b = -\frac{1}{2} \sqrt{b} \Gamma^{-1/2} \quad (2.15)$$

then

$$B = \sqrt{b} \Gamma^{-1/2} \quad (2.16)$$

non-dispersive medium, and the reason is not hard to see. A Lagrangian principle of the form $\delta \int f ds = 0$ can be expected to hold only if two arbitrary events (or at least events arbitrary between limits) can be joined by an extremal, or equivalently if the rays from an event fill a 4-dimensional region in space-time. That is not the case for a non-dispersive medium. Only a cone is filled, as indicated in (11.8). It is easy to reconcile (11.8), which uses the local rest frame, with the more general formula (14.18): in the local rest frame we have $\mu_p = 0$, $\mu_4 = 1$, and (14.18) may be written in the following equivalent form:

$$\begin{aligned}
 -q \alpha_4^2 &= 1 + q, \\
 q \alpha_4^2 &= (1+q) \alpha_r \alpha_r, \\
 (1+q) \alpha_p \alpha_p + \alpha_4^2 &= 0, \\
 n^2 \alpha_p \alpha_p + \alpha_4^2 &= 0,
 \end{aligned}
 \tag{14.19}$$

the last of which is the same as (11.8) in a different notation.

We see that the Lagrangian or medium function does not exist for a non-dispersive medium, and even in the case of a dispersive medium its determination in explicit form is in general not possible. It is in fact better to discuss rays by means of the Hamiltonian equations (9.5) instead of trying to make use of the Euler-Lagrange equations (13.14); the equations (9.3) can be used for non-dispersive media as well as for dispersive media.

central value from the Dirac delta. In this case, for a single isotropic dispersion, the central, zero-effective index as in (15.11), the non-relativistic stationary principle is $\int \delta \mathcal{L} = -\alpha \int \delta \mathcal{L} = 0$ and \mathcal{L} as in (15.21), \mathcal{L} being given by (15.22) with $\mathcal{E} = \omega$ in (15.11).

The stationary principle can also be written

$$\int \frac{\partial \mathcal{L} / \partial \mathbf{A}^2}{1 \mathcal{L}^2} \mathcal{L}^2 \delta \mathbf{A} = 0, \quad (15.45)$$

If the action is to exhibit minima, the non-relativistic case, under $\mathcal{L}^2 = 0$, $\mathcal{L}^2 = 1$, everywhere, then the stationary principle becomes

$$\int \frac{\partial \mathcal{L} / \partial \mathbf{A}^2}{1 \mathcal{L}^2} \mathcal{L}^2 \delta \mathbf{A} = 0, \quad (15.46)$$

If, further, the action is homogeneous, so that \mathcal{L} and \mathcal{L}^2 are constants, we have

$$\int \frac{\partial \mathcal{L} / \partial \mathbf{A}^2}{1 \mathcal{L}^2} \mathcal{L}^2 \delta \mathbf{A} = 0, \quad (15.47)$$

no matter how much \mathcal{L} and \mathcal{L}^2 depend on time and position. This stationary principle should be compared with Fermi's principle for waves, the frequency of each wave not being the real part of the dispersion relation. In fact, (15.47) is a stationary principle $\int \delta \mathcal{L} = 0$, whereas Fermi's principle is the analogue of Fermi's stationary principle, usually written in differential form as

$$\int \delta \mathcal{L} = 0.$$

It is clear to see the Klein-Gordon equation (21.11) has the same structure as Fermi's principle (21.11) for waves, the real part of the dispersion relation being replaced by the real part of the dispersion relation. This is similar to what is stated in (21.11) for the

$x^{1/2}$ solve the equation to the point and from (1) the given result. Then

$$\frac{d(x^{1/2})}{dx} = \frac{1}{2x^{1/2}}, \quad (15.1)$$

so that

$$\begin{aligned} \frac{1}{2} x^{-1/2} + 1 + 2 - 1 &= 1 + 2 + 2x^{1/2}, \\ 1 + 2 &= 1 + 1 + 2x^{1/2}, \end{aligned} \quad (15.2)$$

and (1) becomes (3) to

$$1 = 1 + 1 + 2 + 2x^{1/2}, \quad (15.3)$$

(15.3) gives the following quadratic equation for x :

$$2x^{1/2}x^{1/2} - 2 - (1+1) = 0. \quad (15.4)$$

Thus, according to problem (15.4),

$$x = \frac{1}{4x^{1/2}} [1 \pm \sqrt{1 + 4x^{1/2}(1+1)}], \quad (15.5)$$

Thus, taking x to be positive, we have

$$\begin{aligned} 2x^{1/2} + 1 &= 1 + 1, \\ 2x^{1/2} &= 1 + 1 + 2x^{1/2} + 2x^{1/2} \\ &= 1 + 1 + 2 + 2x^{1/2}, \end{aligned} \quad (15.6)$$

and (15.6) gives

$$1 + 1 + 1 = \frac{d(x^{1/2})}{dx} \left(\frac{d(x^{1/2})}{dx} = \frac{1}{2x^{1/2}} \right). \quad (15.7)$$

Remember that x is given in terms of $(x^{1/2})$ by (15.5), so that we have the

$$\begin{aligned} \mu_1 + \sum_{\alpha} \mu_{\alpha} \gamma_{\alpha} &= -\mu_2, & \mu_2 &= \sum_{\alpha} \mu_{\alpha} \gamma_{\alpha} / \alpha = \mu_1, \\ \mu_3 &= 0, & \mu_4 &= \pm \mu_5. \end{aligned} \quad (15.3)$$

We shall discuss the solution to the simultaneous set of linear (quasi-linear) differential equations valid for non-zero frequencies and later explain, in the (7.5), the +/minus-frequency equation in

$$\partial^2 \psi(\sigma, x) = (\sigma^2) \psi(\sigma, x) = 0, \quad (15.4)$$

where $\psi(\sigma, x) = \psi(\sigma^2, x)$ is a constant. For the case where the equations (15.3)

$$\frac{\partial \mu_1}{\partial \sigma} = -\frac{\partial \mu_2}{\partial \sigma} = \mu_1 - \mu_2 (\sigma^2),$$

$$\frac{\partial \mu_3}{\partial \sigma} = -\frac{\partial \mu_4}{\partial \sigma} = \mu_3 (\sigma^2) \mu_5 \mu_6,$$

we have another partial derivative ($\mu_{\alpha, \sigma} = \partial \mu_{\alpha} / \partial \sigma$):

$$\frac{\partial \mu_1}{\partial \sigma} = \mu_1 - \mu_2 (\sigma^2), \quad \frac{\partial \mu_2}{\partial \sigma} = \mu_2 (\sigma^2) (\mu_3 \mu_4 + \mu_5 \mu_6),$$

$$\frac{\partial \mu_3}{\partial \sigma} = \mu_3 - \mu_4 (\sigma^2), \quad \frac{\partial \mu_4}{\partial \sigma} = \mu_4 (\sigma^2) (\mu_3 \mu_5 + \mu_6),$$

$$\frac{\partial \mu_5}{\partial \sigma} = \mu_5, \quad \frac{\partial \mu_6}{\partial \sigma} = 0, \quad (15.5)$$

$$\frac{\partial \mu_7}{\partial \sigma} = \mu_7 - \mu_8 (\sigma^2), \quad \frac{\partial \mu_8}{\partial \sigma} = 0.$$

Now let us set $\mu_7 = 0$:

$$\sigma^2 \mu_8 = \mu_1 \mu_2 + \mu_3 \mu_4 + \mu_5 \mu_6. \quad (15.6)$$

We note the first integral

substituting for \mathbf{E} from (15.10), we can re-express the force acting on the \mathbf{Q}^{\pm} via familiar expressions, introduced in QM.5:

$$\mathbf{F}^{\pm} = \frac{(\omega \mathbf{M}^{\pm})^2 + \gamma \omega \mathbf{M}^{\pm}}{(\omega \mu)^2} \quad (15.11)$$

4. Spin Hall of Light

In analogy of the situation set up by equation (15.10) above, the passage of light through a cylinder of glass exhibits some interesting properties in the case of Section 7. An incoming monochromatic wave (in a rotating fluid medium) incident in the xz plane, that is, in the plane of symmetry, is split into two waves in the xy plane (15.11).

Consider, then, a PDM (i.e. wave) incident along the x -axis with constant angular velocity ω so that the frequency is

$$\begin{aligned} \omega_1 &= -\omega \omega_2, & \omega_2 &= \omega \omega_1, & \omega_3 &= \omega^2, \\ \omega^2 &= \omega_1^2 \omega_2^2, & \omega^2 &= \omega_1^2 + \omega_2^2. \end{aligned} \quad (15.12)$$

Setting

$$\begin{aligned} \mathbf{E}_1 &= (1 - \omega^2 \mu^2)^{-1/2} = (1 - \omega^2 \mu^2 / \omega^2)^{-1/2}, & (15.13) \\ &+ i\tau = \omega \mu \tau / \omega = \end{aligned}$$

we have for the x -velocity μ_x of the medium

$$\frac{\partial \mathcal{L}}{\partial \mathbf{y}} = \sigma_1^2 + \sigma_2^2 + \mathbf{e} + \lambda (\mathbf{D}^T \mathbf{y}) \mathbf{e} \quad (15.10)$$

$$\frac{\partial \mathcal{L}}{\partial \lambda} = \mathbf{e} \quad (15.11)$$

$$\mathbf{y} = -\sigma_1 \mathbf{e}_1 - \sigma_2 \mathbf{e}_2 = -\text{ones}() \quad (15.12)$$

With this as a starting solution, the ADMM routine converges to the next of iteration:

By (15.12) we have

$$\sigma_1^2 = \sigma_2^2 = \sigma_3^2 = \sigma_4^2 = \lambda (\mathbf{D}^T \mathbf{y})^T \mathbf{e} = \lambda \quad (15.13)$$

and so (15.14) may be written

$$\begin{aligned} \frac{\partial \mathcal{L}}{\partial \mathbf{x}} &= -\sigma_1^2 \mathbf{e}_1 - \sigma_2^2 \mathbf{e}_2 + \lambda (\sigma_1 \mathbf{e}_1^T + \sigma_2 \mathbf{e}_2^T) \mathbf{e} \\ &= -\sigma_1^2 \mathbf{e}_1 - \sigma_2^2 \mathbf{e}_2 + \lambda (\mathbf{e}_1 \mathbf{e}_1^T + \sigma_2 \mathbf{e}_2 \mathbf{e}_2^T) (\mathbf{e}_1 + \mathbf{e}_2) \mathbf{e} = \sigma_2^2 \mathbf{e}_2 \mathbf{e} \end{aligned} \quad (15.14)$$

Note that the right hand side consists of constants and some functions of λ .

To find a stationary (15.14)

$$\frac{\partial \mathcal{L}}{\partial \lambda} (\lambda \mathbf{e}^T) = \mathbf{e}_1 \frac{\partial \sigma_1^2}{\partial \lambda} + \mathbf{e}_2 \frac{\partial \sigma_2^2}{\partial \lambda} = \mathbf{e} \quad (15.15)$$

and so (15.15) gives a differential equation of the second order in iterates λ as a function of \mathbf{e} :

$$\frac{\partial^2}{\partial \lambda^2} (\lambda \mathbf{e}^T) = -\sigma_1^2 \mathbf{e}_1 - \sigma_2^2 \mathbf{e}_2 + \lambda (\mathbf{e}_1 \mathbf{e}_1^T + \sigma_2 \mathbf{e}_2 \mathbf{e}_2^T) (\mathbf{e}_1 + \mathbf{e}_2) \mathbf{e} = \sigma_2^2 \mathbf{e}_2 \mathbf{e} \quad (15.16)$$

in which

$$\sigma_1^2 = \text{constant}, \quad \sigma_2^2 = \text{constant}, \quad (15.4)$$

in accord of which we may treat σ_1 and σ_2 as constants. The only condition of interest will be the second [cf. (15.12)].

Let us define \mathcal{F} and \mathcal{G} by

$$\mathcal{F} = \sigma_1^2 x_1 + \sigma_2^2 x_2 = 0, \quad \mathcal{G} = \sigma_2^2 x_1 + \sigma_1^2 x_2 = 0 \quad (15.5)$$

then, by (15.11)

$$\begin{aligned} \sigma_1^2 p_{11} + \sigma_1^2 p_{21} + \sigma_2^2 p_{12} + \sigma_2^2 p_{22} &= 0 & \sigma_1^2 p_{11} + \sigma_1^2 p_{21} + \sigma_2^2 p_{12} + \sigma_2^2 p_{22} &= -\sigma_1^2 \sigma_2^2 \\ \sigma_1^2 p_{11} &= -\sigma_1^2 \sigma_2^2 & \sigma_2^2 p_{11} &= -\sigma_1^2 \sigma_2^2 \end{aligned} \quad (15.6)$$

and by (15.11) with $\sigma^2 = \sigma_1^2$ obtain

$$\begin{aligned} p_{1,2}^2 &= -\sigma_1^2 \sigma_2^2 \sigma_2^2 / \sigma_1^2 & p_{1,2}^2 &= -\sigma_1^2 \sigma_2^2 / \sigma_1^2 \\ p_{2,1}^2 &= \sigma_1^2 \sigma_2^2 / \sigma_2^2 & p_{2,1}^2 &= \sigma_1^2 \sigma_2^2 / \sigma_2^2 \end{aligned} \quad (15.7)$$

so that

$$\begin{aligned} \sigma_1^2 p_{1,2}^2 + \sigma_2^2 p_{2,1}^2 &= -\sigma_1^2 \sigma_2^2 \sigma_2^2 / \sigma_1^2 \\ \sigma_1^2 p_{2,1}^2 + \sigma_2^2 p_{1,2}^2 &= \sigma_1^2 \sigma_2^2 \sigma_2^2 / \sigma_2^2 \end{aligned} \quad (15.8)$$

and

$$\begin{aligned} \sigma_1^2 p_{1,2}^2 + \sigma_2^2 p_{2,1}^2 &= -\sigma_1^2 \sigma_2^2 \\ \sigma_1^2 p_{2,1}^2 + \sigma_2^2 p_{1,2}^2 &= -\sigma_1^2 \sigma_2^2 \end{aligned} \quad (15.9)$$

By use of the table \mathcal{F} and \mathcal{G} as given in (15.5), substitute for the derivatives of $\sigma_1 = \sigma_2 = \sigma_1^2 = \sigma_2^2$ from (15.6), and sum up, cf. (15.8), (15.9), (15.11). This gives

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$$\begin{aligned}
 \frac{d\tilde{y}}{dt} &= \tilde{y} - \lambda + \beta\lambda = \sigma_1 \tilde{y} \\
 &= \tilde{y} - \lambda \left(\frac{\partial}{\partial \lambda} \tilde{y} \right) = \sigma_2 (\lambda - \beta) \\
 &= \tilde{y} - \frac{\lambda}{\alpha} \frac{\alpha + \sigma_2 (\lambda - \beta)}{\lambda + \omega^2 \beta^2 / \alpha^2},
 \end{aligned} \tag{18.20}$$

Since λ has been fixed as a function of \tilde{y} from (18.15), from (18.20) given \tilde{y} as a function of \tilde{y} . Thus the path of the system is determined. As the time dependence is fixed by (18.6)

$$\begin{aligned}
 \frac{d\tilde{y}}{dt} &= \sigma_2 - \lambda \tilde{y} (\beta + \alpha) \\
 &= \sigma_2 \left[1 + \frac{\alpha(\lambda - \beta)}{\lambda + \omega^2 \beta^2 / \alpha^2} \right],
 \end{aligned} \tag{18.21}$$

Thus it is easy to get a fixed integral of (18.20) (multiply by $\lambda + \omega^2 \beta^2 / \alpha^2$). The capital structure of the firm is essentially reduced to quadrature.

The formulas are much simplified if the top starts from the axis of rotation. The angular velocity of the planet about this axis being then zero. Let us take as initial conditions: The $\tilde{y} = 0$,

$$\begin{aligned}
 \lambda_1 &= \lambda_2 = \lambda_3 = \lambda_4 = \lambda, \\
 \alpha &= \beta, \quad \omega^2 \beta_3 = \omega \beta_4 =
 \end{aligned} \tag{18.22}$$

β_4 being the inclination of the initial top to the axis of rotation. By (18.21), (18.22) we have

$$\tilde{y} = 0, \quad \lambda = \beta, \tag{18.23}$$

$$h(t) = \omega \sqrt{g_0} / v, \quad v^2 = v_0^2 + Y_0^2, \quad Y_0 = Q - \omega^2 r^2 / v^2 r^2. \quad (15.11)$$

Using (15.11), we have

$$\omega_0 = \sigma_0 + \frac{1}{2} \quad (15.12)$$

and σ_0 is a constant, so one of our three independent variables from λ to Y_0 is reduced to:

Let us write (15.12) with multiplication: from (15.1) we find

$$h = v^2 r + \omega \sqrt{g_0} / v. \quad (15.13)$$

and if we define a new dimensionless constant k by

$$k = - \frac{\omega \sqrt{g_0}}{v^2 r}, \quad (15.14)$$

our equation for r becomes

$$\frac{d}{dt} (k r^2) = -\sigma_0^2 - r^2 - \frac{\omega \sqrt{g_0} (Q - R)}{1 - \omega^2 r^2 / v^2 r^2} r - \frac{R}{1 - \omega^2 r^2 / v^2 r^2} r. \quad (15.15)$$

and, by (15.11),

$$\frac{d}{dt} (k r^2) = -1 - \frac{\sigma_0^2}{\sigma_0^2} \left\{ 1 + \frac{R(Q-R)}{1 - \omega^2 r^2 / v^2 r^2} r - \frac{R}{1 - \omega^2 r^2 / v^2 r^2} r \right\}. \quad (15.16)$$

Integration of the RHS is a constant except r :

So we introduce the arbitrary angle θ by

$$r_1 = r \cos \theta, \quad r_2 = r \sin \theta, \quad \tan \theta = r_2 / r_1. \quad (15.17)$$

and write by (15.1)

$$r^2 \frac{d\theta}{dt} = r_2 \frac{dr_2}{dt} - r_1 \frac{dr_1}{dt} = R - (\sigma_0 r^2 + r^2). \quad (15.18)$$

Thus

$$= \frac{R_1}{R_2} = \frac{1}{2} \tan \theta_2 = \frac{1}{2} \frac{R_2}{R_1}, \quad (15-41)$$

so that

$$\cot \theta = 2 = \frac{R_1}{R_2} \tan \theta_2 = \theta_2, \quad (15-42)$$

If we choose the axis so that $\theta = 0$, according to the ray optic discussion,

This gives us the angular loci of the ray as shown behind the surface, with the z as parallel to θ_2 .

So for all calculations you want, just use the equivalent looking θ_2 ($\theta = 0$), assuming the direction is (15-42), or θ .

$$1 \frac{R_1}{R_2} \theta^2 = \tan^2 \theta_2 = \theta^2 \theta^2, \quad (15-43)$$

Thus

$$\theta = \frac{1}{2} \tan \theta_2 \sqrt{\frac{R_2}{R_1}} = \frac{1}{2} \tan \theta_2 \sqrt{\frac{R_2 - z^2}{R_1}}, \quad (15-44)$$

Thus

$$= \theta = \frac{1}{2} \tan \theta_2 \tan 2\theta_2 = \quad (15-45)$$

or

$$\theta = \tan \theta_2 \theta_2 = \frac{1}{2} \tan \theta_2 = \theta^2 \theta_2^2, \quad (15-46)$$

so the ratio θ^2 / θ_2^2 indicates. The first term corresponds to paraxial

approximation; the second term indicates a measure of the ray angle from the axis of rotation. To the same order (15-46) gives the the ratio of bending of the ray

$$= \frac{\theta_1}{\theta_2} = \frac{1}{2} \frac{R_2}{R_1} \tan \theta_2, \quad (15-47)$$

and hence

$$\theta = \frac{1}{2} \frac{R_2}{R_1} \tan \theta_2 = \theta_2, \quad (15-48)$$

and the equation (15.10) for θ simplifies to

$$\frac{d}{dt} \left(\frac{1}{\sin^2 \theta} \right) = -2 + \frac{2}{\sin^2 \theta} \left(1 - \frac{2}{1 - \cos^2 \theta / r^2} \right), \quad (15.11)$$

leading (15.11) to give

$$\tan^2 \theta = -2 - \frac{2}{\sin^2 \theta}, \quad \frac{d\theta}{dt} = -\frac{1}{2} \sec \theta, \quad (15.12)$$

and so (15.11) can be written

$$\frac{d}{dt} \left(\frac{1}{\sin^2 \theta} \right) = \frac{1}{2} \sec^2 \theta = 1 + \frac{2 \sec^2 \theta}{1 - \cos^2 \theta / r^2}, \quad (15.13)$$

multiplying by $(1/r^2) / \sin^2 \theta$ and integrating to get

$$\left(\frac{1}{\sin^2 \theta} \right) \left(\frac{1}{r^2} \right) = \frac{1}{r^2} \left(\frac{1}{\sin^2 \theta} - 1 \right) - \frac{2}{r^2 \sin^2 \theta} \log \left(1 - \frac{\cos^2 \theta}{r^2} \right), \quad (15.14)$$

which, if we define θ to

$$\tan \theta = \cos \theta / r, \quad (15.15)$$

can be expressed as

$$\tan^2 \theta = \frac{1}{2} \sec^2 \theta = 1 - \frac{2}{r^2} \frac{\sec^2 \theta}{1 - \cos^2 \theta / r^2} \log \left(1 - \frac{\cos^2 \theta}{r^2} \right). \quad (15.16)$$

As for $\dot{\theta}$ and \dot{r} , we substitute from (15.15) and (15.16) in (15.12) and (15.13).

NOTICE

$$\frac{d\theta}{dt} = \frac{d\theta}{dr} \frac{dr}{dt} = \frac{\sec \theta}{r}, \quad (15.17)$$

$$\frac{1}{2} \frac{dr}{dt} = \frac{1}{2} \sec \theta \left(1 + \frac{2}{1 - \cos^2 \theta / r^2} \right). \quad (15.18)$$

remains an \mathcal{O}_T -the "ideal" in $\tilde{\chi}$, the usual Hamiltonian function in $\tilde{\Omega}$ is then restricted to elements of the "type", and as (17.13) passes the first integral

$$\tilde{\Omega}(\mathcal{O}_T, \gamma) = \text{const.} \quad (17.14)$$

This constant then acts here as radius.

If $\tilde{\Omega}(\mathcal{O}_T, x) = 0$ happens to be the stationary-frequency equation of some optical system, then the condition of (17.13) provides us with the rays, as previously obtained. As these rays are directed in the same sense of direction of (17.13), obtained by our shift with arbitrary typical values of $x_p = \sigma_p$, and subject to $\tilde{\Delta}(\mathcal{O}_T, x) = 2$. This clear situation can be of advantage in discussing systems of rays, because it allows us more directly to deal with classical theory. For example, we may wish to come to a (pre-)image of two transitions, in which the conditions are $x_p = \sigma_p + \tilde{\chi} + \tilde{\Omega}_p$ and then we accept as

$$\oint (\sigma_p dx_p + \tilde{\Omega} d\tilde{\chi}) \quad (17.15)$$

as our relative integral function which is usually written

$$\oint (p_x dx_x + p dt) \quad (17.16)$$

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$$\dot{\phi} = \frac{\partial \mathcal{L}}{\partial \dot{\phi}} = \frac{\partial}{\partial \dot{\phi}} \text{const } \dot{\phi}_0 + \dot{\phi} \quad (18.6)$$

It is clear that the argument of the present section leads with the propagation of light in an isotropic medium rotating about an axis with constant angular velocity, the medium being non-dispersive and of constant properties so that the refractive index is a constant. The theory is general to (18.5) because from the consideration of rays emanating from a point in the axis of rotation, we finally are approximated to this case by assuming $\dot{\phi} / \dot{\phi}_0 \rightarrow 1$ and $\dot{\phi} \rightarrow \dot{\phi}_0$.

15. A photon in a mechanical system with four degrees of freedom

In the discussion of Section 14, a photon appears as a mechanical system with three degrees of freedom: coordinates x, y, z , momenta $\sigma_x, \sigma_y, \sigma_z$ (boundary lines, and σ_z boundary energy (if we put $\sigma = h \cdot \nu = \dot{\phi}$). But there is another way of looking at this extended system.*

Let $\mathcal{L}(x, y, z, \dot{x}, \dot{y}, \dot{z}, \sigma_x, \sigma_y, \sigma_z, \dot{\phi})$ be any given function of the eight variables: $\sigma_x, \sigma_y, \sigma_z, \dot{\phi}$. Recall the equations

$$\frac{\partial \mathcal{L}}{\partial x} = \frac{\partial \mathcal{L}}{\partial x} = \frac{\partial \mathcal{L}}{\partial \dot{x}} = \dots = \frac{\partial \mathcal{L}}{\partial \dot{x}} \quad (19.1)$$

to recognize these as the equations of motion of a photon in classical form for a mechanical system with three degrees of freedom. (The coordinates are x, y, z .)

* Suggested by Professor G. Abraham in a letter.

