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## THEOREMS ON GENERALIZED DEDEKIND SUMS

T. M. Apostol

1. Introduction. Generalized Dedekind sums $s_{p}(h, k)$, defined by
(1)

$$
s_{p}(h, k)=\sum_{\mu=1}^{k-1} \frac{\mu}{k} \bar{B}_{P}\left(\frac{h \mu}{k}\right)=\sum_{\mu=1}^{k-1} \frac{\mu}{k} B_{p}\left(\frac{h \mu}{k}-\left[\frac{h \mu}{k}\right]\right),
$$

were introduced by the author [1]. The integers $h$ and $k$ are assumed relatively prime, $B_{p}(x)$ is the $p$-th Bernoulli function, $B_{p}(x)$ the $p$-th Bernoulli polynomial (for definitions see $[1 ;(2.11),(2.12)]$ ), and $[x]$ is the greatest integer $\leq x$. For even values of the integer $p$ the sums (1) are trivial (see [1; (4.13)]) and we assume in what follows that $p$ is odd. These sums enjoy a reciprocity law, namely;

$$
(p+1)\left(h k^{p} s_{p}(h, k)+k h^{p} s_{p}(k, h)\right)
$$

(2)

$$
=p B_{p+1}+\sum_{s=0}^{p+1}\binom{p+1}{s}(-1)^{s} B_{s} B_{p+1-s} h^{s} k^{p+1-s} .
$$

The $B$ 's being Bernoulli numbers*. An arithmetic proof of (2) is given in [1] by a method closely related to a general summation technique recently developed by Mordell [5]. When $p=1$, the sums

$$
\begin{equation*}
s_{1}(h, k)=\sum_{\mu=1}^{k-1} \frac{\mu}{k}\left(\frac{h_{\mu}}{k}-\left[\frac{h_{\mu}}{k}\right]-\frac{1}{2}\right) \tag{3}
\end{equation*}
$$

are known as Dedekind sums and are usually denoted by $s(h, k)$. Aside from being of interest from an arithmetical standpoint, these sums also occur in the asymptotic theory of partitions and have been studied in a large number of papers, for example [1], [3], [5], [6], [7], [8], [9], [10], and [11].

In this paper we establish a connection between the sums (1) and certain finite sums involving Hurwitz zeta functions which makes it possible to give a short analytic proof of (2).

[^0]2. The Hurwitz zeta function and Dedekind sums. The Hurwitz zeta function $\zeta(s, a)$ is defined for $\Re(s)>1$ and $a \neq 0,-1,-2, \cdots$, by the series
$$
\zeta(s, a)=\sum_{n=0}^{\infty}(n+a)^{-s}
$$
and its relation to $s_{p}(h, k)$ is given in the following theorem.
Theorem 1. For odd $p>1$ we have
\[

$$
\begin{equation*}
s_{p}(h, k)=i p!(2 \pi i k)^{-p} \sum_{\mu=1}^{k-1} \cot \frac{\pi h \mu}{k} \zeta\left(p, \frac{\mu}{k}\right) \tag{4}
\end{equation*}
$$

\]

while for $p=1$ we have the two equivalent expressions

$$
\begin{equation*}
s(h, k)=\frac{1}{4 k} \sum_{\mu=1}^{k-1} \cot \frac{\pi h \mu}{k} \cot \frac{\pi \mu}{k} \tag{5}
\end{equation*}
$$

and

$$
\begin{equation*}
s(h, k)=\frac{-1}{2 \pi k} \sum_{\mu=1}^{k-1} \cot \frac{\pi h \mu}{k} \frac{\Gamma^{\prime}(\mu / k)}{\Gamma(\mu / k)} \tag{6}
\end{equation*}
$$

Formula (5) is due to Rademacher [8], who derived it from the Fourier series expansion of (3). We will give here a purely arithmetic proof of (5) based on finite rather than infinite Fourier series. Secondly, we establish the equivalence of (5) and (6) and then prove (4). Finally, we indicate how (5) and (6) can be thought of as limiting cases of (4).

Proof of (5): The function $\bar{B}_{1}(x)$ is given by

$$
\bar{B}_{1}(x)= \begin{cases}x-[x]-1 / 2 & \text { if } x \neq \text { integer } \\ 0 & \text { otherwise }\end{cases}
$$

Therefore, by formula (2.5) of [10] we may write

$$
\begin{equation*}
s(h, k)=\sum_{\mu \bmod k} \bar{B}_{1}(\mu / k) \bar{B}_{1}(h \mu / k) \tag{7}
\end{equation*}
$$

From Eisenstein's finite Fourier series expansion [4; p.318] we have

$$
\begin{equation*}
\bar{B}_{1}(h \mu / k)=-\frac{1}{2 k} \sum_{\nu=1}^{k-1} \sin \frac{2 \pi h \nu \mu}{k} \cot \frac{\pi \nu}{k} . \tag{8}
\end{equation*}
$$

Using (8) in each factor of the summand in (7), we obtain
(9) $s(h, k)=\frac{1}{8 k^{2}} \sum_{\lambda=1}^{k-1} \sum_{\nu=1}^{k-1} \cot \frac{\pi \nu}{k} \cot \frac{\pi \lambda}{k} \sum_{\mu \bmod k}\left(\cos \frac{2 \pi \mu(\lambda-\nu h)}{k}\right.$

$$
\left.-\cos \frac{2 \pi \mu(\lambda+\nu h)}{k}\right)
$$

because of the identity $2 \sin x \sin y=\cos (x-y)-\cos (x+y)$. Since we have

$$
\sum_{\mu \bmod k} \cos \frac{2 \pi \mu(\lambda \pm \nu h)}{k}= \begin{cases}k & \text { if } \lambda \pm \nu h \equiv 0(\bmod k) \\ 0 & \text { otherwise }\end{cases}
$$

for each fixed $\nu$ only one value of $\lambda$ gives a nonzero contribution to each sum in the second member of (9), namely $\lambda \equiv \nu h(\bmod k)$ in the first sum and $\lambda \equiv-\nu h$ $(\bmod k)$ in the second. Therefore we have

$$
s(h, k)=\frac{1}{8 k} \sum_{\nu=1}^{k-1} \cot \frac{\pi \nu}{k} \cot \frac{\pi h \nu}{k}-\frac{1}{8 k} \sum_{\nu=1}^{k-1} \cot \frac{\pi \nu}{k} \cot \frac{-\pi h \nu}{k}
$$

and this is the same as (5).
Proof that (5) and (6) are equivalent: The relation [2; p. 163]

$$
\begin{equation*}
\frac{\Gamma^{\prime}(\mu / k)}{\Gamma(\mu / k)}=-\gamma-\log k-\frac{\pi}{2} \cot \frac{\pi \mu}{k} \tag{10}
\end{equation*}
$$

$$
+\sum_{n \leq k / 2}^{\prime} \cos \frac{2 \pi n \mu}{k} \log \left(2-2 \cos \frac{2 \pi n}{k}\right)
$$

where $\gamma$ is Euler's constant and the prime indicates that when $k$ is even the last term is to be multiplied by $1 / 2$, is due to Gauss. Multiplying both sides of (10) by $\cot (\pi h \mu / k)$ and summing on $\mu$ shows the equivalence of (5) and (6) upon observing that we have

$$
\sum_{\mu=1}^{k-1} f(\mu)=0
$$

whenever $f$ is an odd function of $\mu$ which is also periodic $\bmod k$.
Proof of (4): Formula (4.11) of [1] gives a representation of $s_{p}(h, k)$ as an infinite series which, with some simplification, can be written in the form

$$
s_{p}(h, k)=i p!(2 \pi i)^{-p} \sum_{\substack{n=1 \\ n \neq 0(\bmod k)}}^{\infty} n^{-p} \cot \frac{n \pi h}{k} .
$$

Writing $n=q k+\mu$, with $q=0,1,2, \cdots, \infty$, and $\mu=1,2, \cdots, k-1$, we obtain

$$
\begin{aligned}
s_{p}(h, k) & =i p!(2 \pi i)^{-p} \sum_{\mu=1}^{k-1} \sum_{q=0}^{\infty}(q k+\mu)^{-p} \cot \frac{\pi h \mu}{k} \\
& =i p!(2 \pi i k)^{-p} \sum_{\mu=1}^{k-1} \cot \frac{\pi h \mu}{k} \zeta(p, \mu / k),
\end{aligned}
$$

where we must assume $p>1$ in order to insure that the series involved should be absolutely convergent and the rearrangements valid. This proves (4). We cannot hope for a proof of (4) along the lines of our proof of (5) because of the nonelementary nature of $\zeta(s, a)$.

If in (4) we replace $p!$ by $\Gamma(p+1)$ and let $p$ be a complex variable which tends to 1 , then we can show that the two expressions for $s(h, k)$ in (5) and (6) occur naturally as limiting cases of the right member of (4). We first observe that, although the function $\zeta(s, a)$ has a pole at $s=1$, the sum

$$
\begin{equation*}
\sum_{\mu=1}^{k-1} \cot \frac{\pi h \mu}{k} \zeta(s, \mu / k) \tag{12}
\end{equation*}
$$

is regular at $s=1$. This is easily seen by using the expansion
$\zeta(s, a)=\frac{1}{s-1}-\frac{\Gamma^{\prime}(a)}{\Gamma(a)}+\mathrm{O}(s-1)$
(as $s \longrightarrow 1$ )
obtained from Whittaker and Watson [12; p. 271], substituting in (12) and using (11) to obtain

$$
\lim _{s \rightarrow 1} \sum_{\mu=1}^{k-1} \cot \frac{\pi h \mu}{k} \zeta(s, \mu / k)=-\sum_{\mu=1}^{k-1} \cot \frac{\pi h \mu}{k} \frac{\Gamma^{\prime}(\mu / k)}{\Gamma(\mu / k)},
$$

which shows that the right member of (4) tends to the right member of (6) as $p \rightarrow 1$.

The connection between (5) and (4) can be obtained by using Hurwitz's functional equation in the form given by Rademacher [6; (1.24)], namely:

$$
\begin{aligned}
& \zeta(s, \mu / k)=2 \Gamma(1-s)(2 \pi k)^{s-1} \sum_{\lambda=1}^{k}\left(\cos \frac{\pi s}{2} \sin \frac{2 \pi \lambda \mu}{k}\right. \\
&\left.+\sin \frac{\pi s}{2} \cos \frac{2 \pi \lambda \mu}{k}\right) \zeta\left(1-s, \frac{\lambda}{k}\right)
\end{aligned}
$$

this being valid for $s \neq 1, \mathrm{l} \leq \mu \leq k$. Multiplying by $\cot (\pi h \mu / k)$, summing on $\mu$ and using (11) leads to

$$
\sum_{\mu=1}^{k-1} \cot \frac{\pi h \mu}{k} \zeta(s, \mu / k)
$$

$$
\begin{equation*}
=2 \Gamma(1-s)(2 \pi k)^{s-1} \cos \frac{\pi s}{2} \sum_{\lambda, \mu=1}^{k-1} \cot \frac{\pi h \mu}{k} \sin \frac{2 \pi \lambda \mu}{k} \zeta(1-s, \lambda / k) . \tag{13}
\end{equation*}
$$

Since $\zeta(0, a)=1 / 2-a$, when $s$ tends to 1 the right member of (13) approaches the value

$$
\begin{aligned}
\frac{1}{2 k} \sum_{\lambda=1}^{k-1} \sum_{\mu=1}^{k-1} \cot \frac{\pi h \mu}{k} \sin & \frac{2 \pi \lambda \mu}{k}\left(\frac{1}{2}-\frac{\lambda}{k}\right) \\
& =\frac{-1}{2 k^{2}} \sum_{\mu=1}^{k-1} \cot \frac{\pi h \mu}{k} \sum_{\lambda=1}^{k-1} \lambda \sin \frac{2 \pi \lambda \mu}{k}
\end{aligned}
$$

because of (11). Noticing that the last sum on $\lambda$ is the imaginary part of the sum

$$
\sum_{\lambda=1}^{k-1} \lambda e^{2 \pi i \lambda \mu / k}=k /\left(e^{2 \pi i \mu / k}-1\right)=-\frac{k i}{2} \cot (\pi \mu / k)-\frac{k}{2},
$$

we see that the right member of (4) also tends to the right member of (5) when $p \rightarrow 1$.
3. Proof of the reciprocity law. We can now give a proof of the reciprocity law (2) using complex integration. This proof is of additional interest in that we use properties of $\zeta(s, a)$ for fixed $s$ and variable $a$. We will need the following facts about $\zeta(s, a)$ :

$$
\begin{equation*}
\zeta(s, a)=\zeta(s, a+1)+a^{-s}, \tag{14}
\end{equation*}
$$

$$
\begin{equation*}
\zeta(s, a+1)=\zeta(s)-s \zeta(s+1) a+\mathrm{O}\left(a^{2}\right) \quad \text { as } a \rightarrow 0 \tag{15}
\end{equation*}
$$

for $0 \leq \Re(a) \leq M$, (M fixed $), \zeta(s, a)$ tends uniformly to 0 as $\Im(a) \rightarrow$ $\pm \infty$. (The uniformity is with respect to $\Re(a))$.

Equation (14) follows at once from the definition of $\zeta(s, a)$ and (15) is merely the beginning of the Taylor series for $\zeta(s, a+1)$ near $a=0$. Here $\zeta(s)=$ $\zeta(s, 1)$ is Riemann's zeta function. Relation (16) can be readily obtained, for example, by applying the Riemann-Lebesgue theorem to the integral representation [2; p. 266]:

$$
\Gamma(s) \zeta(s, a)=\int_{0}^{\infty} \frac{t^{s-1} e^{-t \Re(a)}}{1-e^{-t}} e^{-i t} \Im(a) d t
$$

valid for $\Re(s)>1$ and $\Re(a)>0$. This gives (16) for $0<\Re(a) \leq M$ and (14) proves it for $0 \leq \Re(a) \leq M$.

Because of (4), the reciprocity formula (2) can now be put into the following form:

Theorem 2. For odd $p>1$ we have

$$
\begin{array}{r}
\frac{i(p+1)!}{(2 \pi i)^{p}}\left\{h \sum_{\mu=1}^{k-1} \cot \frac{\pi h \mu}{k} \zeta(p, \mu / k)+k \sum_{\nu=1}^{h-1} \cot \frac{\pi k \nu}{h} \zeta(p, \nu / h)\right\}  \tag{17}\\
=p B_{p+1}+\sum_{s=0}^{p+1}\binom{p+1}{s} B_{s} B_{p+1-s} h^{s} k^{p+1-s} .
\end{array}
$$

Proof. We apply Cauchy's residue theorem to the function

$$
f(z)=\cot \pi z \cot (\pi h z / k) \quad \zeta(p, z / k)
$$

Integrating in the positive sense around a contour $C$ consisting of a rectangle whose vertices are the points $\pm i T, k \pm i T$, with small semi-circular detours $C_{0}$ and $C_{k}$ around the points $z=0, z=k$, traversed along the $\operatorname{arcs} z=\epsilon e^{i \theta}$ and $z=k+\epsilon e^{i \theta}$, respectively, where $\pi / 2 \leq \theta \leq 3 \pi / 2$, and $0<\epsilon<1 / h$. Ultimately, $\epsilon$ will tend to 0 and $T>1 / 2$ will tend to $\infty$. The integrand $f(z)$ has first order poles at the points $z=1,2, \cdots, k-1$ due to the factor cot $\pi z$, and at the points $z=k / h, 2 k / h, \cdots,(h-1) k / h \quad$ because of the factor $\cot (\pi h z / k)$. By (14) we have

$$
\zeta(p, z / k)=\zeta(p, z / k+1)+(k / z)^{p},
$$

so that the point $z=0$ is a pole of order $p+2$ for $f(z)$. Using the power series expansion

$$
\begin{equation*}
\pi z \cot \pi z=\sum_{n=0}^{\infty} \frac{(2 \pi i)^{n} B_{n}}{n!} z^{n} \tag{18}
\end{equation*}
$$

in the neighborhood of $z=0$ (with the understanding that $B_{1}$ should be replaced by 0), and (15) with $a=z / k$ we find that Cauchy's theorem gives us

$$
\frac{1}{2 \pi i} \int_{C} f(z) d z=\frac{1}{\pi} \sum_{\mu=1}^{k-1} \cot \frac{\pi h \mu}{k} \zeta(p, \mu / k)
$$

$$
\begin{equation*}
+\frac{k}{\pi h} \sum_{\nu=1}^{h-1} \cot \frac{\pi k \nu}{h} \zeta(p, \nu / h)-\frac{p}{\pi^{2} h} \zeta(p+1)+R_{0} \tag{19}
\end{equation*}
$$

where

$$
R_{0}=\operatorname{Res}_{z=0} \cot \pi z \cot (\pi h z / k)(k / z)^{p} .
$$

We now observe that by periodicity of the cotangent and by (14), the contribution to the integral from the part of $C$ consisting of vertical line segments is

$$
\left(\int_{i T}^{i \epsilon}+\mathcal{S}_{-i \epsilon}^{-i T}\right) \cot \pi z \quad \cot (\pi h z / k)(k / z)^{p} d z
$$

and this vanishes since the integrand is an odd function of $z$. Next, the integrals along the horizontal segments tend to zero as $T \rightarrow \infty$ since, for $0 \leq x \leq k$ we have cot $\pi(x+i y) \rightarrow \mp i$ and, by (16), $\zeta(p,(x+i y) / k)$ tends to 0 uniformly
in $x$ as $y \rightarrow \pm \infty$. Finally, combining the integrals over $C_{0}$ and $C_{k}$ by means of (14) and letting $T \rightarrow \infty$ we obtain

$$
\lim _{T \rightarrow \infty} \int_{C} f(z) d z=\int_{C_{0}} \cot \pi z \quad \cot (\pi h z / k)(k / z)^{p} d z
$$

When $\epsilon \longrightarrow 0$ we find

$$
\lim _{\epsilon \rightarrow 0} \quad \int_{C_{0}}=\pi i R_{0}
$$

so that equation (19) leads to the result

$$
\begin{equation*}
\frac{1}{\pi} \sum_{\mu=1}^{k-1} \cot \frac{\pi h \mu}{k} \zeta(p, \mu / k)+\frac{k}{\pi h} \sum_{\nu=1}^{h-1} \cot \frac{\pi k \nu}{h} \zeta(p, \nu / h) \tag{20}
\end{equation*}
$$

$$
=\frac{p}{\pi^{2} h} \zeta(p+1)-\frac{1}{2} R_{0} .
$$

From (18) we easily calculate that

$$
R_{0}=\frac{2 i(2 \pi i)^{p}}{\pi h(p+1)!} \sum_{s=0}^{p+1}\binom{p+1}{s} B_{s} B_{p+1-s} h^{s} k^{p+1-s}
$$

and, since we have

$$
\zeta(p+1)=-\frac{(2 \pi i)^{p+1} B_{p+1}}{2(p+1)!}
$$

equation (20) yields (17) and the proof is complete.
In [8], Rademacher gives a proof for the case $p=1$ using (5) instead of (4). Apparently unaware of [8], K. Iseki [3] has given a proof very much like Rademacher's analytic proof for the case $p=1$ in a recent paper.

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## ADDENDUM TO 'ON THE LERCH ZETA FUNCTION'

T. M. Apostol

Professor L. Carlitz has been kind enough to point out that the functions $\beta_{n}(a, \alpha)$ which were used in [1] to evaluate the Lerch zeta function $\phi(x, a, s)$ for negative integer values of $s$ have occurred elsewhere in the literature in other connections, for example in [2] and [3]. As Carlitz points out, formula (3.3) of [1] leads to the result

$$
\alpha^{m} \beta_{n}(m, \alpha)-\beta_{n}(0, \alpha)=n \sum_{a=0}^{m-1} a^{n-1} \alpha^{a}
$$

which, for integer values of the variable $a$, makes apparent the relation of the functions $\beta_{n}(a, \alpha)$ with the Mirimanoff polynomials discussed by Vandiver in [3].

There is a misprint in the next to last equation on p . 164 of [1]. The coefficient of $a^{2} / 2$ in the expression for $\phi(x, a,-2)$ should read $i$ cot $\pi x+1$ instead of $i \cot \pi x+1 / 4$.

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## EXTENSION OF FUNCTIONS ON FULLY NORMAL SPACES

Richard Arens

1. Introduction. Starting from the recent discovery of A. H. Stone that metric spaces are "paracompact" [12] (paracompactness means that every open covering has a refinement only a finite number of whose members meet a suitable neighborhood of each point [5]), J. Dugundji has been able to extend to metric spaces certain techniques in the theory of retracts which were hitherto applicable at most to separable metric spaces [6]. The cornerstone of his method is a theorem (see 2.4, below) according to which a continuous function on a closed set $A$ of a metric space $X$ with values in a convex ( $=$ "locally convex") topological linear space $L$ may be extended to the whole space $X$, indeed without enlarging the convex hull of the image. Essentially, the possibility of doing this for a locally separable metric space $X$ is implicitly given by a procedure for the real valued case in [10].

One of the problems to which we address ourselves in this paper is that of determining whether the assumption that $X$ is metric can be reduced to $X$ is merely paracompact. The answer (see 6, below) is no. However, we have fairly general results which imply that if $L$ is metric and complete (and $X$ is paracompact) then the extension is possible (4.1, below). Our proof utilizes a process of extending a pseudo-metric on $A$ to all of $X$, which is ultimately based on a theorem of Hausdorff. We generalize Hausdorff's theorem (3.2 and 3.4) and incidentally show how Dugundji's result enables one to construct a short proof of Hausdorff's theorem.

None of these extension theorems can properly be regarded as a true generalization of Tietze's extension theorem, which deals with mappings on normal spaces with values on the line or in the Hilbert cube, since there exist normal, not fully normal spaces. In order to provide a generalization of Tietze's theorem, we have shown by way of application that the Hilbert cube may be replaced by any compact convex subset of a normed linear space (4.3).
S. Kakutani [10] has introduced the notion of "simultaneous extension regarded as a linear positive operation," in his case of real valued functions on a locally separable metric space $X$ : this means that it is possible so to extend
all continuous functions such that $(\lambda f+g)^{e}=\lambda f^{e}+g^{e}$ and $f^{e} \geq 0$ whenever $f \geq 0$, where the superscribed $e$ indicates the extension. We show that this is possible in the more general case in which $X$ is metric and $L$ convex, the order preserving feature of Kakutani's formulation being naturally reformulated as a nonenlargement of the convex hull of the image (2.6). What is perhaps more surprising is that the "simultaneous extension," while possible in more general cases (4.2), is not possible under as general conditions as those under which the individual extension (as in 4.1) is possible. As a matter of fact, we tie up the notion of simultaneous linear and order relation preserving extension with that of "sweeping" a measure from $X$ onto $A$, and thus show that it is not always possible even for compact Hausdorff spaces $X$ (6.1).

The question arises whether among the simultaneous extensions which preserve linear and order relations, there are not some which preserve quadratic polynomial relations as well. It has already been shown by Yoshizawa just when this is possible: at least when $X$ is compact, $A$ must be a retract of $X$.

We have inserted a section (5) showing that the "simultaneous extension" for real valued functions can be derived from the "individual" extension of a suitable continuous function with values in an infinite dimensional space, as well as from the fundamental Lemma (2.1) directly.

In formulating our results, we shall always speak of "fully normal" [13] rather than "paracompact" spaces, although it is known that these two properties coincide [12]. We do this because we use the full normality as such, using Stone's result only for the metric case.
2. Extension of functions on metric space. One of the main geometric ideas underlying the process of extension involved here is contained in the following construction (cf. [6, 4.3]).
2.1. Lemma. Let $X$ be a metric space, and $A$ a closed subset of $X$. Then there exists a family $g_{V}$ of continuous real-valued functions defined on $X$, and a similarly indexed family of points $a_{V}$ of $A$ such that
2.2. each $g_{V}$ vanishes on $A$, all but a finite number of the $g_{V}$ vanish in some neighborhood of each point of $X$, the sum $\sum_{V} g_{V}(x)=1$ for all $x$ in $X-A$ and each $g_{V}(x)$ is nonnegative;
2.3. for each $a$ in $A$ and each $V$, if $g_{V}(x)>0$ then $d\left(a, a_{V}\right)<3 d(a, x)$ (d is the metric in $X$ ) and $d(a, x)<d\left(a, a_{V}\right)+2 d(x, A)$.

Proof. To each $x$ in $X-A$, assign the open sphere of radius $d(x, A) / 4$. Since $X-A$ is metric it is paracompact [12], so that this covering of $X-A$ has a
refinement $R$ which is "neighborhood finite" (see [12]). For each $V$ in $R$ there is a point $x$ such that $V$ is contained in the open sphere about $x$ of radius $d(x, A) / 4$; select such a point and call it $x_{V}$. Also select $a_{V}$ in $A$ such that $d\left(a_{V}, x_{V}\right)$ $<(5 / 4) d\left(x_{V}, A\right)$. Let $f_{V}(x)=d(x, X-V)$ for every $x$ in $X$. Since $R$ is neighborhood finite, each $x$ in $X$ has a neighborhood on which all but finitely many $f_{V}$ vanish, and so $s(x)=\sum_{V} \in{ }_{R} f_{V}(x)$ is both finite and continuous. Since $s(x)$ is never 0 , the functions $g_{V}=f_{V} / s$ are continuous. It is easy to see that they provide the kind of "Dieudonné partition" required by 2.2.

We now turn to 2.3. Suppose $g_{V}(x)>0$. Then $x$ belongs to the open set $V$. From what has been said about $V, x_{V}$, and $a_{V}$ follows

$$
d\left(a_{V}, x\right) \leq d\left(a_{V}, x_{V}\right)+d\left(x_{V}, x\right)<(5 / 4) d\left(x_{V}, A\right)+(1 / 4) d\left(x_{V}, A\right) .
$$

Of course

$$
\begin{aligned}
d\left(x_{V}, A\right) & \leq d\left(x_{V}, a\right) \\
& \leq d\left(x_{V}, x\right)+d(x, a)<(1 / 4) d\left(x_{V}, A\right)+d(x, a),
\end{aligned}
$$

and so (3/4) $d\left(x_{V}, A\right)<d(x, a)$. Thus $d\left(a_{V}, x\right)<(6 / 4) d\left(x_{V}, A\right)<2 d(x, a)$. Thus finally we have half of 2.3 , since

$$
d\left(a, a_{V}\right) \leq d(a, x)+d\left(x, a_{V}\right)<d(a, x)+2 d(x, a)=3 d(a, x)
$$

For the second half of 2.3 , we note first that $d(x, A)>3 / 4 d\left(x_{V}, A\right)$. On the other hand,

$$
d\left(a_{V}, x\right) \leq d\left(a_{V}, x_{V}\right)+d\left(x_{V}, x\right)<(6 / 4) d\left(x_{V}, A\right),
$$

which is thus less than $2 d(x, A)$. Finally,

$$
d(a, x) \leq d\left(a, a_{V}\right)+d\left(a_{V}, x\right)<d\left(a, a_{V}\right)+2 d(x, A)
$$

Thus the proof of 2.3 is complete.
Geometrically, the lemma given above says that $X-A$ can be so mapped into the finite dimensional faces of the "formal simplex" with vertices equal to the points of $A$, in such a way that as $x$ tends to a point $a_{0}$ of $A$, the vertices of the carrier of the image of $x$ all tend to $a_{0}$, in the topology of $A$. With this picture in mind, it is easy to imagine how functions on $A$ with values in a convex set $E$, can be extended to all of $X$. The next result $[6,4.1]$ makes this precise.
2.4. Theorem. Let $K$ be a convex subset of a convex topological linear space $L$ (cf. [14]). Let A be a closed subset of a metric space X. Let f be continuous on $A$ with values in $K$. Then $f$ may be extended continuously to a function $f^{e}$ defined on $X$ with values in $K$.

Proof. Using the result and notation of 2.1, we define $f^{e}$ at once by setting $f^{e}(x)=\sum_{V} g_{V}(x) f\left(a_{V}\right)$, for $x$ in $X-A$, and $f^{e}(a)=f(a)$ for $a$ in $A$. There remains only the proof of continuity. Now the topology of $L$ (and thus $K$ ) can be based on neighborhoods of 0 defined by relations $\|k\|<1$ where $\|\cdots\|$ is one of the pseudo-norms of $L$, according to von Neumann's idea (cf. [14]). Select any point $a$ of $A$. There is a positive $r$ such that $d(a, b)<r$ implies $\| f(a)$ $-f(b) \|<1$ for $b$ in $A$. Now suppose $d(a, x)<r / 3$. For those finitely many $V$ for which $g_{V}(x)$ is not 0 , we have $d\left(a, a_{V}\right)<r$, so that

$$
\left\|f^{e}(x)-f(a)\right\| \leq \sum g_{V}(x)\left\|f\left(a_{V}\right)-f(a)\right\|<1
$$

This shows the continuity of $f^{e}$ at any point of $A$. At points $x$ of $X-A$ we can find a neighborhood in which only finitely many $g_{V}$ do not vanish, so that $f^{e}$ is continuous there also. The rest of 2.4 is obvious. The second half of 2.3 is not needed for this proof.

The fact that a single formula, so to speak, can be chosen to perform the extension can be expressed in several ways. Suppose $K_{1}$ and $K_{2}$ are convex subsets of two convex topological linear spaces, and let there be an affine mapping $m$ of $K_{1}$ into $K_{2}$. Suppose $f_{1}, f_{2}$ are functions as in 2.4 with values in $K_{1}, K_{2}$ respectively, but satisfying the condition $m\left(f_{1}(a)\right)=f_{2}(a)$ for all $a$ in $A$. If we use the same system $g_{V}, a_{V}$ in extending $f_{2}$ as in extending $f_{1}$ then we surely obtain $m\left(f_{1}^{e}(x)\right)=f_{2}^{e}(x)$ for all $x$ in $X$. We shall abbreviate this by saying that the process of extension when applied to all possible $f$ is consistent, and note the result:
2.5. Theorem. Each $f$ satisfying the hypothesis of 2.4 with $K$ variable but $A$ and $X$ constant may be so extended that the entire process is consistent.

Another kind of consistency or simultaneity is expressed as follows.
2.6. Theorem. Let $K$ be a (linear or possibly merely convex) subset of a convex topological linear space, and let $A$ be a closed subset of the metric space $X$. Let $F$ be the class of continuous functions on $A$ with values in $K$. Then each $f$ may be extended by an $f^{e}$ (using 2.4) in such a way that, for $f_{1}, \ldots, f_{n}$ in $F$ and $c_{1}, \ldots, c_{n}$ real numbers (nonnegative with sum 1 when $K$ is merely
convex), we have

$$
\left(c_{1} f_{1}+\cdots+c_{n} f_{n}\right)^{e}=c_{1} f_{1}^{e}+\cdots+c_{n} f_{n}^{e}
$$

This result is a generalization of Kakutani's theorem [10] on "simultaneous extension of continuous functions considered as a positive linear operation". The only real advance of 2.6 over Kakutani's theorem is the removal of separability, although Kakutani limits $F$ to the space of bounded real valued continuous functions $C(A)$.

An addendum to 2.4 and 2.6 is of interest:
2.7. Under the conditions of 2.4 or 2.6 , if there is an $f$ and a point a of $A$ such that $f$ is constant on a neighborhood (relative to $A$ ) of $a$, then $f^{e}$ is constant on a neighborhood (relative to X) of $a$.

In fact, suppose $f\left(a^{\prime}\right)$ is constant for $d\left(a, a^{\prime}\right) \leq 3 e$ and $a^{\prime}$ in $A$. Then $f^{e}(x)$ is constant for $d(a, x) \leq e$, since then $d\left(a, a_{V}\right) \leq 3 e$.
3. Extension of pseudo-metrics. Let $X$ be a topological space. Let $s$ be a real-valued function of two variables defined in $X$ such that

$$
s(y, x)=s(x, y) \geq 0, s(x, z) \leq s(x, y)+s(y, z)
$$

and such that the set of $x$ such that $s(x, y)<e$ is open for each $e>0$ and $y$. Then $s$ is a pseudo-metric. It falls short of being a metric in that $s\left(x_{n}, y\right) \longrightarrow 0$ $\left[s(x, y)=0\right.$ ] does not necessarily imply $x_{n} \longrightarrow y(x=y)$. Our first result is in the direction of an extension of a pseudo-metric from a closed set to the whole space.
3.1. Lemma. Let $X$ be a fully normal [13] topological space, and let $q$ be a pseudo-metric defined on a closed subset $A$ of $X$. Then there is a pseudo-metric $s$ defined in all of $X$ such that for $x, y$ in $A$ and $k=4,5, \ldots$, if $s(x, y)<2^{-k}$ then $q(x, y)<2^{-k}$.

Proof. Select a positive integer $n$. Construct an open covering $U$ consisting of those open sets $V$ which intersect $A$ in a set of $q$-diameter less than $2^{-n}$. Using the terminology, notation, and results of [13] we obtain $U \stackrel{*}{>} U_{1} \stackrel{*}{>} U_{2} \stackrel{*}{>} \ldots$ [13, V-7.4], and a pseudo-metric $r_{n}$ such that [13, V-7.5, correcting $\in$ to $\notin$ ] $x \notin S\left(y, U_{p}\right)$ implies $r_{n}(x, y)>2^{-(p+2)}$ for $p=1,2, \ldots$. We may also assume $r_{n}(x, y) \leq 1$. We can thus form $r(x, y)=\sum_{n} 2^{-n} r_{n}(x, y)$. This is clearly a pseudo-metric. Suppose $r(x, y)<2^{-k}$, for $k \geq 4$ and $x, y$ in $A$. Then $r_{k-3}(x, y)<$ $2^{-3}$. Hence $x \in S\left(y, U_{1}\right)$, this covering being the one obtained for $n=k-3$.
(We have omitted an index showing dependence on $n$.) Since $U_{1} \stackrel{*}{<} U$ we obtain $q(x, y)<2^{-k+3}$. Setting $s=2^{-3} r$ gives the required pseudo-metric, completing the proof of 3.1.

We remind the reader that compact Hausdorff spaces and metric spaces are fully normal [13, V-8.14, VI-4.5].

This lemma is actually all we need in order to extend the results of $\delta 2$ as we shall do later. However, by an application of a theorem of Hausdorff, we can improve 3.1 aesthetically by obtaining a pseudo-metric $s$ which is an extension to $X$ of the original $q$. In fact, rather than refer to Hausdorff's theorem, we first give a new proof since it is an interesting application of 2.1 , is much shorter than Hausdorff's, and shows in passing how a metric space may be isometrically imbedded in its own space of bounded continuous functions (cf. [4, p.187]). The present proof resembles that in [11] more than that in [9]. However, Kuratowski's proof, besides requiring separability, generally does not provide an isometric, but merely topological imbedding (see below, and also [9, p.47]).
3.2. Theorem. [Hausdorff]. Let $A$ be a closed subset of a metric space $X$, and let $f$ be a continuous mapping of $A$ into another metric space $B$. Then $B$ can be isometrically imbedded in a metric space $Y$ such that $f$ can be continuously extended to $X$ with values in $Y$, such that $f$ is a homeomorphism of $X-A$ with $Y-B$, and such that $B$ is closed in $Y$.

Proof: For any space $S$ let $C(S)$ denote the Banach space of real-valued continuous bounded functions $g$ on $S$, with $\|g\|=\sup _{x} \in S|g(x)|$.

To begin the proof, obtain for $X$ a bounded metric $d$. The metric $r$ in $B$ we must not alter, of course. For $b$ in $B$, let $r_{b}$ denote the function with values $r_{b}\left(b^{\prime}\right)=r\left(b, b^{\prime}\right)$. This function is not necessarily bounded, but $r_{b}-r_{c}$ is bounded (cf. [4, p.187]) and $\left\|r_{b}-r_{c}\right\|=r(b, c)$, where $b, c$ are points of $B$. Select a point $o$ in $A$, to be held constant. The function $\phi$ defined for $a$ in $A$ by $\phi(a)=r_{f(a)}-r_{f(o)}$ evidently maps $A$ into $C(A)$. Indeed, since

$$
\left\|\phi(a)-\phi\left(a^{\prime}\right)\right\|=\left\|r_{f(a)}-r_{f\left(a^{\prime}\right)}\right\|=r\left(f(a), f\left(a^{\prime}\right)\right),
$$

the map $\phi$ is continuous. It may therefore by extended to all of $X$ by 2.4 , and we denote the extension also be $\phi$. Now form $L=C(B) \times R \times C(X)$, where

$$
\|(h, j, k)\|=\max (\|h\|,|j|,\|k\|)
$$

and $R$ is the real number system. For $x, y$ in $X$ define $d_{x}(y)=d(x, y)$ as earlier, and let $d(x)=d(x, A)$. For $x$ in $X$, define $F(x)=\left[\phi(x), d(x), d(x) d_{x}\right]$ in $L$.

This $F$ is obviously continuous. Define $B_{1}=F(A)$ and $Y=F(X)$, both subsets of $L$. Clearly $B_{1}$ is closed relative to $Y$. Now for each $a$ in $A$ we obtain $f(a)$ on one hand and $F(a)$ on the other. We now show that it sets up an isometry between $B$ and $B_{1}$. In fact,

$$
\left\|F(a)-F\left(a^{\prime}\right)\right\|=\left\|\phi(a)-\phi\left(a^{\prime}\right)\right\|=\left\|r_{f(a)}-r_{f\left(a^{\prime}\right)}\right\|=r\left[f(a), f\left(a^{\prime}\right)\right]
$$

as mentioned earlier. If we identify $B$ with $B_{1}$, then $F$ becomes an extension of $f$, continuous on all of $X$. Suppose $F(x)=F(y)$, where $y$ belongs to $X-A$. Then $d(x)=d(y)>0$; hence $d_{x}=d_{y}$, which means $x=y$. Thus $F$ has an inverse on $Y-B_{1}$. We shall show that it is an homeomorphism. Let $y \in X-A$ and suppose $F(x) \longrightarrow F(y)$. Then $d(x) \longrightarrow d(y)>0$, and $d(x) d_{x} \longrightarrow d(y) d_{y}$. From this we conclude $d_{x} \rightarrow d_{y}$ or $d(x, y)=\left\|d_{x}-d_{y}\right\| \rightarrow 0$. Thus 3.2 , Hausdorff's theorem, is proved. It is to be borne in mind that it was not known in 1938 that metric spaces were paracompact.

We go on to establish a refinement of 3.2 also due to Hausdorff.
3.3. Theorem. If the $f$ in 3.2 is a homeomorphism of $A$ with $B$ then it can be arranged that $F$ also is a homeomorphism.

To establish 3.3, Hausdorff [9, p.46] modifies the construction of $F$. It is an interesting fact that the $F$ we construct automatically satisfies 3.3. The only nontrivial part of the proof of 3.3 is that if $F(x) \longrightarrow F(a)$ for $x$ in $X-A$ and $a$ in $A$, then $x \rightarrow a$ in X. Therefore, suppose $F(x) \longrightarrow F(a)$. Let

$$
h=g(x)-g(a)=\sum g_{V}(x)\left[r_{f\left(a_{V}\right)}-r_{f(a)}\right]
$$

where the $g_{V}$ and $a_{V}$ are described in 2.1. Now $|h(f(a))| \leq\|h\| \longrightarrow 0$. But

$$
h[f(a)]=\sum g_{V}(x) r\left[f\left(a_{V}\right), f(a)\right]
$$

is not less than the least of those $r\left[f\left(a_{V}\right), f(a)\right]$ which appear in this sum, that is, for which $g_{V}(x)$ is not 0 . Denote the $a_{V}$ in question by $a_{W}$, where of course $W$ depends on $x$. Since $r\left(f\left(a_{W}\right), f(a)\right]$ tends to 0 and $f$ is a homeomorphism on $A$, we see that $a_{W} \longrightarrow a$. From 2.3 we obtain $d(a, x)<d\left(a, a_{W}\right)+2 d(x)$, and so $x \longrightarrow a$, as desired.

These two results have the following consequence.
3.4. Corollary. Let $A$ be a closed subset of a metric space X. Let r be a pseudo-metric defined on $A$. Then this pseudo-metric may be extended to all of $X$ in such a way that
3.41 in $X-A$ it is equivalent to the metric $d$ of $X$;
3.42 in $x \in X-A$ then for some positive $e, r(x, y)<e$ implies $y \in X-A$;
3.43 if $r$ is a metric equivalent to $d$ on $A$, the extension is equivalent to $d$ on all of $X$.

Proof. In $A$, form the equivalence classes for the relation $r(x, y)=0$, and metrize in the obvious way using $r$. Call the resulting space $B$. The natural mapping of $A$ onto $B$ satisfies the hypothesis of 3.2 . Let $m$ be the metric in $Y$. Then $m[F(x), F(y)]$ gives the desired extension of $r(x, y)$.

We can now provide the finishing touch to 3.1.
3.5. Theorem. Let $X$ be a fully normal topological space, and let $q$ be a pseudo-metric defined on a closed subset of $X$. Then $q$ can be extended to be a pseudo-metric on $X$.

The proof is based on 3.1 and 3.4 as follows. Using the $s$ of 3.1, partition $X$ into a set $X^{*}$ of equivalence classes according to the relation $s(x, y)=0$, denoting the class containing $x$ by $x^{*}$, and so on. Define $s^{*}\left(x^{*}, y^{*}\right)=s(x, y)$; this is a valid definition, which makes $X^{*}$ a metric space, and the natural mapping of $X$ onto $X^{*}$ is continuous. Let $A^{*}$ be the closure in $X^{*}$ of the image of $A$. The conclusion of 3.1 shows that $q$ may be carried over in unique fashion to $A^{*}$, to form a pseudo-metric $q^{*}$. An appeal to 3.4 extends $q^{*}$ to $X^{*}$, and $q(x, y)=q^{*}\left(x^{*}\right.$, $\left.y^{*}\right)$ provides the desired extension.

Note that we have no use for 3.41-3.43 in 3.5 because the $s$ was not given to us in advance.
4. Extension of functions on fully normal spaces. In the next result, the metric for $X$ in 2.4 is shifted to $K$.
4.1. Theorem. Let $A$ be a closed subset of a fully normal space $X$. Let $f$ be continuous on $A$ with values in a complete convex metric subset $K$ of a convex topological linear space L. Then $f$ can be continuously extended to $X$ with all values still in $K$.

Proof. In $A$ define the pseudo-metric $q\left(a, a^{\prime}\right)=m\left[f(a), f\left(a^{\prime}\right)\right]$, where $m$ is the metric, and extend $q$ to $X$ by 3.5. Let $A_{0}$ be the set of $x$ such that $q(x, A)$ $=0$. Given $e>0$ and $x$ in $A_{0}$, let $S_{e}$ be the set of $a$ in $A$ such that $q(x, a)<e$. The $f\left(S_{e}\right)$ form a nested system in $K$, and their diameters shrink to 0 . Hence there is just one point, which we call $f(x)$, common to all. This provides an
extension of $f$ to $A_{0}$. Now partition $X$ into a set $X^{*}$ of equivalence classes under the relation $q(x, y)=0$, denoting the class containing $x$ by $x^{*}$, and so on. Define $q^{*}\left(x^{*}, y^{*}\right)=q(x, y)$; this makes $X^{*}$ into a metric space and the continuous natural image of $X$. In this natural mapping, $A_{0}$ passes onto a closed subset $A^{*}$ of $X^{*}$. The function $f^{*}\left(a^{*}\right)=f(a), a$ in $A_{0}$, is continuous (indeed isometric) on $A^{*}$. It can be extended to all of $X^{*}$, by 2.4. Going back and defining $f(x)=$ $f^{*}\left(x^{*}\right)$, we get an extension of $f$ with the desired properties.

We shall show in 6 that a "simultaneous extension" of the type of 2.6 cannot always be obtained if the hypothesis is merely that of 4.1 for each of the functions involved. However, using the procedure of 4.1 and the result of 2.6 , the reader may prove the following:
4.2. Theorem. Let $A$ be a closed subset of a fully normal space $X$. Let $F$ be a linear (convex) set of functions each defined on $A$ and with values in a complete metric linear (convex) subset $K$ of a convex topological linear space $L$. Furthermore let there be defined on A a pseudo-metric $q$ such that for each $f$ in $F$ and for each positive $r$ there is a positive $s$ such that $q\left(a, a^{\prime}\right)<s$ implies $m\left[f(a), f\left(a^{\prime}\right)\right]<r_{s}$ where $m$ is the metric in K. Then a simultaneous extension (in the sense of 2.6) can be made for all the $f$ in $F$.

None of the preceeding results can properly be claimed to be a generalization of Tietze's extension theorem, since we always require more than normality of $X$. We do not know whether the following is true: if $A$ is a closed subset of a normal space $X$, and $f$ maps $A$ continuously into a bounded closed convex subset $K$ of a Banach space $L$, then $f$ can be continuously extended to $X$ with values in $K$. Of course, in the finite dimensional case of $L$, this result is an easy consequence of the original theorem. In this case, we can replace "bounded" by "compact", and in this form the theorem does admit generalization.
4.3. Theorem. Let $A$ be a closed subset of a normal space $X$. Let $K$ be a compact convex subset of a normed linear space L. Let $f$ be a continuous function on $A$ with values in $K$. Then $f$ can be continuously extended to $X$ with values in $K$ (see note added in proof).

Proof: Since $K$ is separable, we can find a countable family $v_{1}, v_{2}, \ldots$ of bounded linear functionals on $L$ such that if $u, u^{\prime}$ belong to $K$ and $u\left(v_{n}\right)=u^{\prime}\left(v_{n}\right)$ for all $n$, then $u=u^{\prime}$. (cf. [2, p.484, "Note"]). We now imbed $K$ in the space (s) of [4]. For $u$ in $K$, define $U(u)_{n}=v_{n}(u)$. This mapping is continuous and one-toone, and hence a homeomorphism. We may therefore forget about the original $L$ and regard $K$ as a compact convex subset of $(s)$. By 2.4 , since ( $s$ ) is metrizable, we can obtain a retraction of $(s)$ on $K$, that is, a continuous $r$ such that $r(u) \in K$
for $u$ in $(s)$, and $r(u)=u$ for $u$ in $K$. Let $f_{n}(a)=f(a)_{n}$, the $n$-th coordinate of $f(a)$ in (s). By Tietze's original theorem, this $f_{n}$ may be extended continuously to all of $X$. Defining $f_{0}(x)=\left[f_{1}(x), f_{2}(x), \ldots\right]$ in $(s)$ we obtain a mapping of $X$ into (s). Setting $f^{-}(x)=r\left[f_{0}(x)\right]$ gives the desired extension.
5. Simultaneous extension of real-valued functions. This section merely shows that special cases of 2.6 and 4.2 in which the linear space is the real number system $R$ (or any finite dimensional linear space) can be reduced to 2.4 or 4.1 , respectively, without further inquiry into the method of extension. In other words the possibility of "simultaneous extension" of real-valued functions is a direct consequence of the possibility of a single extension of a function with values in a suitable infinite dimensional space. This sounds quite plausible, but it is perhaps surprising that we must consider conjugate spaces.

Consider first a closed subset $A$ of a metric space $X$, and the spaces $C(A)$ and $C(X)$ of continuous real-valued functions on them. Let $L$ be $C(A)^{-}$, the conjugate space, with the weak topology (see [14], for example). Let $K$ be the set of $\xi$ in $L$ with norm not exceeding 1 and with $\xi \geq 0$ (that is, $\xi(f) \geq 0$ for $f \geq 0$ ). For $a$ in $A$, define $F(a)$ in $K$ by $F(a)(f)=f(a)$. This $F$ is continuous since we are using the weak topology, and $K$ is convex. By 2.4 this $F$ can be extended to $X$. For $f$ in $C(A)$, define $f^{e}$ by $f^{e}(x)=F(x)(f)$. We leave to the reader the completion of the proof of the following:
5.1. Theorem. The operation $f \rightarrow f^{e}$ is a linear, isometric, nonnegative transformation of $C(A)$ into $C(X)$, and $f^{e}$ is an extension of $f$.

In the next section we shall show that 5.1 cannot be generalized for nonmetric $X$ even if $X$ is compact. However, the following is true.
5.2. Theorem. Let $A$ be a closed subset of a fully normal space $X$. Let $S$ be a separable (in the norm topology) closed linear subspace of $C(A)$. Then there is a linear isometric nonnegative transformation $f \longrightarrow f^{e}$ of $S$ into $C(X)$ such that $f^{e}$ is an extension of $f$.

The proof is just like that of 5.1 , except that we appeal to 4.1. To do this we must observe that since $S$ is separable, $K$ in $S^{-}$with the weak topology is metrizable (as is well known), for example with the metric

$$
m(\xi, \eta)=\sum 2^{-n}\left|(\xi-\eta)\left(f_{n}\right)\right|
$$

where the $f_{n}$ are dense in the unit ball of $S$; and that $K$ is compact (AlaogluBourbaki [1]) and thus complete.
6. Applications to measure in topological spaces. Let $X$ be a topological space with a measure, and let $A$ be a subset such that every function of a fixed linear set $F$ of real-valued functions on $A$ can be extended to a summable function on $X$ by a positive linear operation $P$. By defining $J(f)=\int P(f)(x) m(d x)$ for $f$ in $F$, we obtain a functional which may sometimes be represented by an integral (cf. [2, 3] or any of the references given there). When this is true, one obtains a measure $m^{\prime}$ on $A$ which is generally not the mere restriction $\left[m^{\prime}(E)\right.$ $=m(E)$ for $E \in A]$ of $m$ to $A$.

Unfortunately we have not been able to apply this process to any situation to obtain measures in $A$ of a class not more easily obtainable by other methods. This is because of the requirement of the existence of a pseudo-metric in 4.2 with the stated properties, or of the separability of $S$ in 5.2 . The interest of the present section lies mainly in the fact that it is shown that one cannot avoid limitations of this sort. For this purpose we present only one of a variety of theorems, and then show why it cannot be generalized.
6.1. The orem. Let $X$ be a fully normal Hausdorff space and let m be a finite Baire measure [8] such that $m(V)=0$ for an open $V$ only if $V$ is void. Let $A$ be a compact subset of $X$. Let $S$ be a separable subset of $C(4)$. Then there exists a strongly regular measure $m^{\prime}$ in $A$ such that all functions in $S$ are measurable and if $f \in S$ and $f \geq 0, f \neq 0$ then

$$
\int_{A} f(a) m^{\prime}(d a)>0
$$

Proof. Let $Q$ be the normed linear algebra generated by $S$ and 1. By [3, 4.4] we can obtain a measure as described such that

$$
\int_{A} f(a) m^{\prime}(d a)=\int_{X} f^{e}(x) m(d x)
$$

The point to observe is that if $f \geq 0, f \neq 0$, then the same thing is true for $f^{e}$, and thus the right integral is positive.

Why can we not ignore the separability of $S$ in 6.1? Let $A_{0}$ by any uncountable discrete set. By adding a "point at infinity" we obtain a compact space $A$. This space $A$ can be imbedded in a cartesian product $X$ of unit intervals. The obvious product measure [8, p. 158 (2)] has the properties needed for 6.1. Let $S$ $=C(A)$, and, forgetting that $S$ is not separable, apply 6.1. The resulting measure would make every one of the points $A_{0}$ have nonzero measure, and so $A$ itself would not be measurable. This shows why the separability of $S$ in 6.1 cannot be ignored; and it also shows that one cannot ignore the pseudo-metric $q$ in 4.2 or the separability of $S$ in 5.2.

Added in proof: We have recently found, and shall soon publish, a stronger form of 4.3 , namely in which "compact" is replaced by "separable".

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## A SHORT PROOF OF PILLAI'S THEOREM ON NORMAL NUMBERS

John E. Maxfield

1. Introduction. The object of this paper is to give a short proof of the Pillai theorem [2] on normal numbers using the Niven-Zuckerman result [1] as a tool.

Definition 1. A number $\sigma$ is simply normal to the base $r$ if, in the expansion to the base $r$ of the fractional part of $\sigma$, we have $\lim _{n \rightarrow \infty} n_{c} / n=1 / r$ for all $c$, where $n_{c}$ is the number of occurrences of the digit $c$ in the first $n$ digits of $\sigma$.

Definition 2. A number $\sigma$ is normal to the base $r$ if $\sigma, r \sigma, r^{2} \sigma, \ldots$ are each simply normal to all the bases $r, r^{2}, r^{3}, \ldots$.

Theorem (Pillai). A necessary and sufficient condition that a number $\sigma$ be normal to the base $r$ is that it be simply normal to the bases $r, r^{2}, r^{3}, \ldots$.
2. Proof. The necessity of the condition follows from the definition of normality.

To prove sufficiency, assume that $\sigma$ is simply normal to the bases $r, r^{2}$, . . . Let $A=\left(a_{1} a_{2} \cdots a_{v}\right)$ be any fixed sequence of digits (to base $r$ ), where $v=h r-s, h>0,0 \leq s<r$; and consider the occurrence of $A$ in $\sigma$. Count the number of occurrences of $A$ in the collection of sequences of length $h r$. There are $s$ digits free after $v$ of the $h r$ digits are fixed. Thus there are $(s+1) r^{s}$ different occurrences of $A$ in these sequences.

For any positive integer $n$, define $f_{n}(A)$ to be the frequency of the occurrences of $A$ in $\sigma$ except in places where $A$ will straddle the middle of sequences of length $2 h 2^{n-1} r$ starting in places congruent to $1\left(\bmod 2 h 2^{n-1} r\right)$, or where $A$ will straddle the middle of sequences of length $4 h 2^{n-1} r$ starting in places congruent to $1\left(\bmod 4 h 2^{n-1} r\right)$, or $\cdots$, or where $A$ will straddle the middle of sequences of length $2^{s} h 2^{n-1} r$ starting in places congruent to $1\left(\bmod 2^{s} h 2^{n-1} r\right)$, and so on.

Certainly $\lim _{n \rightarrow \infty} f_{n}(A)$, if it exists, will be equal to $f(A)$, the frequency of $A$ in $\sigma$.

We have

$$
f_{1}(A)=\frac{(s+1) r^{s}}{h r r^{h r}}=\frac{1}{r^{v}}-\frac{v-1}{h r^{v+1}},
$$

since there are $h r$ digits of $\sigma$ to base $r$ in each digit of $\sigma$ to base $r^{h r}$, and $\sigma$ is simply normal to the base $r^{h r}$. The number of occurrences of $A$ straddling the middle of blocks of length $2 h r$ is $(v-1) r^{2 h r}+s$. The frequency of these in $\sigma$, where the sequence of length $2 h r$ starts in a place congruent to $1(\bmod 2 h r)$, is

$$
\frac{(v-1) r^{2 h r+s}}{2 h r r^{2 h r}}=\frac{v-1}{2 h r^{v+1}},
$$

since there are $2 h r$ digits of $\sigma$ to base $r$ to each digit of $\sigma$ to base $r^{2 h r}$. Thus

$$
f_{2}(A)=\frac{1}{r^{v}}-\frac{v-1}{h r^{v+1}}+\frac{v-1}{2 h r^{v+1}} .
$$

Similarly,

$$
f_{3}(A)=f_{2}(A)+\frac{v-1}{4 h r^{v+1}}=\frac{1}{r^{v}}-\frac{v-1}{h r^{v+1}}+\frac{v-1}{h r^{v+1}}\left[\frac{1}{2}+\frac{1}{4}\right]
$$

and

$$
f_{n}(A)=\frac{1}{r^{v}}-\frac{v-1}{h r^{v+1}}+\frac{v-1}{h r^{v+1}} \sum_{i=1}^{n-1} 1 / 2^{i}
$$

It follows that

$$
\lim _{n \rightarrow \infty} f_{n}(A)=1 / r^{v}
$$

Accordingly, by the Niven-Zuckerman result [1], stating that a necessary and sufficient condition in order that a number $\sigma$ be normal is that every fixed sequence of $v$ digits occur in the expansion of $\sigma$ with the frequency $1 / r^{v}$, we see that $\sigma$ is normal to the scale $r$.

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# QUASI-CONVEXITY AND THE LOWER SEMICONTINUITY OF MULTIPLE INTEGRALS 

Charles B. Morrey, Jf.

1. Introduction. We are concerned in this paper with integrals of the form

$$
\begin{equation*}
I(z, D)=\int_{D} f\left[x, z^{i}(x), z_{x^{\alpha}}^{i}(x)\right] d x, \tag{1.1}
\end{equation*}
$$

where

$$
\begin{aligned}
& x=\left(x^{1}, \cdots, x^{\nu}\right), z=\left(z^{1}, \cdots, z^{N}\right), p=p_{\alpha}^{i} \\
& \\
& \quad(i=1, \cdots, N ; \alpha=1, \cdots, \nu),
\end{aligned}
$$

$f(x, z, p)$ is continuous in its arguments, and $D$ is a bounded domain.
The object of the paper is to discuss necessary and sufficient conditions on the function $f$ for the integral $I$ to be lower semicontinuous with respect to various types of convergence of the vector functions $z$. Because of the success of the "direct methods" in the Calculus of Variations, many writers have shown that certain integrals are lower semicontinuous. However, the writer knows of no paper in which a necessary condition for lower semicontinuity was discussed, although such a condition is very easy to obtain (see Theorem 2.1).

In $\S_{2}$, a general condition called "quasi-convexity" (see Definition 2.2) on the behavior of $f$ as a function of $p$ is obtained which is both necessary and sufficient for the lower semicontinuity of $I$ with respect to the type of convergence given in Definition 2.1. This condition is that any linear function furnish the absolute minimum to $I(z, D)$ among all Lipschitzian (see below) functions which coincide with it on $D^{*}, D$ being any bounded domain and $D^{*}$ its boundary; here, of course, we consider $f$ to be a function of $p$ only. Section 3 discusses cases involving more general types of convergence and gives an existence theorem. In $\delta 4$, it is shown that if $f(p)$ is continuous and quasi-convex, then it satisfies a certain generalized Weierstrass condition which reduces to the ordinary one (for the case at hand) when $f$ is of class $C^{\prime}$; this is, in turn, seen to be equivalent to the Legendre-Hadamard condition (see (4.8)) (quasi-regularity in its general form) when $f$ is of class $C$. In $\oint 5$, a general sufficient

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condition for quasi-convexity is proved and the necessary condition of $\S 4$ is seen to be sufficient when $f$ is either a quadratic form in the $p_{\alpha}^{i}$ or is the integrand of a parametric problem with $N=\nu+1$. The view of Terpstra's negative result [5] that even the strong Legendre-Hadamard condition ( $>0$ ) does not necessarily imply the existence of an alternating form $C_{i j}^{\alpha \beta} p_{\alpha}^{i} p_{\beta}^{j}\left(C_{i j}^{\beta \alpha}=-C_{i j}^{\alpha \beta}\right.$, and so on) such that

$$
\begin{equation*}
f(p)+C_{i j}^{\alpha \beta} p_{\alpha}^{i} p_{\beta}^{j} \equiv\left(a_{i j}^{\alpha \beta}+C_{i j}^{\alpha \beta}\right) p_{\alpha}^{i} p_{\beta}^{j} \tag{1.2}
\end{equation*}
$$

is positive definite when $\nu>2$, it would seem that there is still a wide gap in the general case between the necessary and sufficient conditions for quasiconvexity which the writer has obtained. In fact, after a great deal of experimentation, the writer is inclined to think that there is no condition of the type discussed, which involves $f$ and only a finite number of its derivatives, and which is both necessary and sufficient for quasi-convexity in the general case.

In (1.2), we have used the usual tensor summation convention, and will continue to use it throughout the paper; unless otherwise specified, the Greek letters will run from 1 to $\nu$ and the Latin letters from 1 to $N$.

We shall denote the sum and difference of vectors of the various sorts $(x, z$, $p$, and so on) in the usual way. We shall define

$$
\left.|x|=\left(x^{\alpha} x^{\alpha}\right)^{1 / 2}, \quad|z|=\left(z^{i} z^{i}\right)\right)^{1 / 2},|p|=\left(p_{\alpha}^{i} p_{\alpha}^{i}\right)^{1 / 2}
$$

If $\zeta(x)$ is a vector function with derivatives, $\pi(x)$ will denote the vector function $\pi_{a}^{i}(x)=\zeta_{x}^{i}(x)$; similar notations involving other letters will be introduced as the occasion demands.

All integrals are Lebesgue integrals, frequently of vector functions. It is sometimes desirable to consider the behavior of a function $z(x)$ with respect to a particular variable $x^{\alpha}$ or to the $\nu-1$ variables $\left(x^{1}, \cdots, x^{\alpha-1}, x^{\alpha+1}\right.$, $\left.\cdots, x^{\nu}\right)$. In such a case, we write $x_{\alpha}^{\prime}$ for $\left(x^{1}, \cdots, x^{\alpha-1}, x^{\alpha+1}, \cdots, x^{\nu}\right)$, $\left(x_{\alpha}^{\prime}, x^{\alpha}\right)$ for $x$ and so on. It is also convenient to write the boundary integrals

$$
\int_{D^{*}} A_{\alpha}(x) d x_{\alpha}^{\prime}
$$

where each $A_{\alpha}(x)$ may be a vector $A_{\alpha}^{i}(x)$ and the boundary $D^{*}$ of the domain is sufficiently regular; such an integral is to be regarded as a Lebesgue-Stieltjes integral with respect to the set function $x_{\alpha}^{\prime}(e)$ on $D^{*}$ chosen so that Green's theorem

$$
\int_{D^{*}} \zeta^{i} d x_{\alpha}^{\prime}=\int_{D} \zeta_{x \alpha}^{i} d x
$$

holds. The closure of a set $E$ will be denoted by $\bar{E}$.
Ordinary functions of class $\Re_{s}, \Re_{s}^{\prime}, \Re_{s^{\prime \prime}}$, and so on, $s \geqq 1$, have been discussed at length in the papers [1] and [2]; the extension to vector functions is trivial. We define the integrals $\bar{D}_{s}(z, G)$ and $D_{s}(z, G)$ by

$$
\bar{D}_{s}(z, G)=\int_{G}|z(x)|^{s} d x+L_{s}(z, G), D_{s}(z, G)=\int_{G}\left[z_{x \alpha}^{i}(x) z_{x a}^{i}(x)\right]^{s / 2} d x
$$

Each function $z$ of class $\Re_{s}$ is equivalent to a function $\bar{z}$ defined uniquely almost everywhere as that number such that the Lebesgue derivative of the set function

$$
\int_{e}\left|z(x)-\bar{z}\left(x_{0}\right)\right|^{s} d x
$$

is zero at $x_{0} ; \bar{z}$ is supposed to be defined at every point $x_{0}$ where such a number exists; $\bar{z}$ is of class $\Re_{s}^{\prime}$ (see [1] and [2]) and is also of class $\Re_{s}^{\prime}$ in any coordinate system related to the original by a regular Lipschitzian transformation (cf. [2], Theorem 6.3; the $\bar{z}$ there used has a slightly different definition from the present one but the present theorem has been proved for vectors $z$ with values in a Riemannian manifold in [4], Lemma 2.3 and Theorem 2.5).

A function $z$ is said to satisfy $a$ (uniform) Lipschitz condition with coefficient $M$ on a set $S$ if and only if

$$
\left|z\left(x_{1}\right)-z\left(x_{2}\right)\right| \leqq M\left|x_{1}-x_{2}\right|, x_{1} \in S, x_{2} \in S .
$$

A function is Lipschitzian if it satisfies a Lipschitz condition.
If $g(y), y=\left(y^{1}, \cdots, y^{n}\right)$, is summable on a domain $D$, we define the $h$ average function $g_{h}$ by

$$
g_{h}(y)=(2 h)^{-n} \int_{y-h}^{y+h} g(\eta) d \eta, \quad h>0 ;
$$

if $g$ is summable then $g_{h}$ is continuous where defined; if $g$ is continuous on $D$ then $g_{h}$ is of class $C^{\prime}$ and $g_{h}$ tends uniformly to $g$ on each bounded closed set interior to $D$; if $g$ is of class $\Re_{s}$ on $D$ then $g_{h}$ tends strongly in $\Re_{s}$ to $g$ on each domain $G$ with $\bar{G} \subset D$ (see [1], Lemma 5.1).

A form

$$
C_{i_{1}, \cdots, i_{\mu}}^{a_{1}, \cdots, a_{\mu}} \pi_{a_{1}}^{i_{1}} \cdots \pi_{a_{\mu}}^{i_{\mu}}\left(\mu \leqq \nu, 1 \leqq \alpha_{\gamma} \leqq \nu, 1 \leqq i_{\gamma} \leqq N, \gamma=1, \cdots, \mu\right),
$$

is called alternating if and only if the $C$ 's satisfy the obvious symmetry requirements and also the antisymmetry condition that

$$
C_{i_{1}, \cdots, i_{\mu}}^{\beta_{1}, \cdots, \beta_{\mu}}= \pm C_{i_{1}, \cdots, i_{\mu}}^{a_{1}, \cdots, a_{\mu}}
$$

according as $\left(\beta_{1}, \cdots, \beta_{\mu}\right)$ is an even or odd permutation of the indices $\left(\alpha_{1}, \cdots, \alpha_{\mu}\right)$; if $\zeta(x)$ is a vector function, then

$$
\mu!C_{i_{1}, \cdots, i_{\mu}}^{a_{1}, \cdots, a_{\mu}} \pi_{a_{1}}^{i_{1}}(x) \cdots \pi_{a_{\mu}}^{i_{\mu}}(x)=C_{i_{1}}^{\alpha_{1}, \cdots, i_{\mu}} \frac{\partial\left(\zeta^{i_{1}}, \cdots, \zeta^{i_{\mu}}\right)}{\partial\left(x^{\alpha_{1}}, \cdots, x^{a_{\mu}}\right)}
$$

the fractions on the right denoting Jacobians.
2. A necessary and sufficient condition for lower-semicontinuity. We begin with some definitions.

Definition 2.1. For the purposes of this section, we say that the vector functions $z_{n}$ tend to the vector function $z$ on the domain $D$ if and only if the $z_{n}$ and $z$ all satisfy a uniform Lipschitz condition on $D$, independent of $n$, and the $z_{n}$ tend uniformly to $z$ on $D$. We shall write $z_{n} \longrightarrow z$ to denote this type of convergence.

Definition 2.2. A function $f\left(p_{\alpha}^{i}\right)$ is said to be quasi-convex if and only if

$$
\int_{D} f[p+\pi(x)] d x \geqq f(p) \cdot m(D), \quad \pi_{\alpha}^{i}(x)=\zeta_{x a}^{i}(x)
$$

for each constant $p$, each domain $D$, and each vector function $\zeta$ which satisfies a uniforn. Lipschitz condition on $D$ and vanishes on $D^{*}$.

We shall show in this section that the integral $I(z, D)$ is lower semicontinuous with respect to the type of convergence specified in Definition 2.1 on each bounded domain $D$ if and only if $f(x, z, p)$ is quasi-convex in $p$ for each fixed $(x, z)$.

Theorem 2.1. Suppose $I(z, D)$ is lower semicontinuous with respect to the type of convergence indicated on every region $D$. Then $f$ is quasi-convex in $p$ for each fixed $(x, z)$.

Proof. Let $x_{0}$ be any point, $R$ be the cell $x_{0} \leqq x^{i} \leqq x_{0}^{i}+h, Q$ be the cell $0 \leqq x^{i} \leqq 1$, and $\zeta$ be any function of class $C^{\prime}$ and periodic in each $x^{i}$ with period 1. Let $z_{0}$ be any function of class $C^{\prime}$ on $R$.

For each $n$, define $\zeta_{n}(x)$ on $R$ by

$$
\zeta_{n}^{i}(x)=n^{-1} h \zeta^{i}\left[n h^{-1}\left(x-x_{0}\right)\right] .
$$

Then

$$
\zeta_{n x^{\alpha}}^{i}(x)=\zeta_{x^{\alpha}}^{i}\left[n h^{-1}\left(x-x_{0}\right)\right]
$$

and

$$
\begin{aligned}
I\left(z_{0}+\right. & \left.\zeta_{n}, R\right)=\int_{R} f\left\{x, z_{0}^{i}(x)+\zeta_{n}^{i}(x), p_{0 \gamma}^{i}(x)+\pi_{\gamma}^{i}\left[n h^{-1}\left(x-x_{0}\right)\right]\right\} d x \\
= & \sum_{\alpha} \int_{R_{\alpha}}\left(f\left\{x, z_{0}^{i}(x)+\zeta_{n}^{i}(x), p_{0 \gamma}^{i}(x)+\pi_{\gamma}^{i}\left[n h^{-1}\left(x-x_{0}\right)\right]\right\}\right. \\
& \left.\quad-f\left\{x_{\alpha}, z_{0}^{i}\left(x_{\alpha}\right), p_{0 \gamma}^{i}\left(x_{\alpha}\right)+\pi_{\gamma}^{i}\left[n h^{-1}\left(x-x_{0}\right)\right]\right\}\right) d x \\
+ & \sum_{\alpha}\left(n^{-1} h\right)^{\nu} \int_{Q} f\left\{x_{\alpha}, z_{0}^{i}\left(x_{\alpha}\right), p_{0 \gamma}^{i}\left(x_{\alpha}\right)+\pi_{\gamma}^{i}(\xi)\right\} d \xi
\end{aligned}
$$

where

$$
\begin{aligned}
& \alpha=\left(\alpha_{1}, \cdots, \alpha_{\nu}\right), R_{\alpha}=R_{\alpha_{1}}, \cdots, \alpha_{\nu}, n^{-1}\left(\alpha_{\beta}-1\right) \leqq x^{\beta} \leqq n^{-1} \alpha_{\beta} \\
& x_{\alpha}=\left(x_{a_{1}, \cdots, a_{\nu}}^{\beta}\right), x_{a_{1}}^{\beta}, \cdots, a_{\nu}=n^{-1}\left(\alpha_{\beta}-1\right), \beta=1, \cdots, \nu
\end{aligned}
$$

As $n \rightarrow \infty$, we see, since $f$ is uniformly continuous on any bounded part of space, $\zeta_{n}(x)$ tends uniformly to zero, and the $\pi_{\gamma}^{i}$ are bounded, that

$$
\lim _{n \rightarrow \infty} I\left(z_{0}+\zeta_{n}, \tilde{K}\right)=\int_{R}\left[\int_{Q} f\left[x, z_{0}(x), p_{0}(x)+\pi(\xi)\right] d \xi\right] d x
$$

From the lower semicontinuity of $I$, we must have

$$
\int_{R}\left\{\int_{Q} f\left[x, z_{0}(x), p_{0}(x)+\pi(\xi)\right] d \xi\right\} d x \geqq \int_{R} f\left[x, z_{0}(x), p_{0}(x)\right] d x
$$

Now, let $x_{0}, z_{0}$, and $p_{0}$ be any constant vectors. By letting

$$
z_{0}(x)=z_{0}+p_{0 \alpha}\left(x^{\alpha}-x_{0}^{\alpha}\right),
$$

dividing by $h^{\nu}$ and letting $h \rightarrow 0$, we obtain

$$
\int_{Q} f\left[x_{0}, z_{0}, p_{0}+\pi(\xi)\right] d \xi \geqq f\left(x_{0}, z_{0}, p_{0}\right)
$$

By approximations, we can extend this to all $\zeta$ which satisfy a uniform Lipschitz condition over the whole space and are periodic of period 1 in each $x^{\alpha}$.

Now, let $D$ be a bounded domain and suppose $\zeta$ satisfies a uniform Lipschitz condition on $D$ and vanishes on $D^{*}$. Let $R$ be a hypercube of edge $h$, with edges parallel to the axes which contains $D$. Extend $\zeta$ to the whole space by first defining it to be zero on $\bar{R}-D$ and then extending it to be periodic of period $h$ in each variable. Then a simple change of function and variable reduces $R$ to $\mathcal{Q}$ and establishes the result.

Lemma 2.1. Suppose $R$ is a cell with edges $\left(2 h^{1}\right), \cdots,\left(2 h^{\nu}\right)$ and center $x_{0}$. Let $h$ be the smallest $h^{a}$. Suppose also that $0<k<h$, that $\zeta^{*}$ satisfies a uniform Lipschitz condition with coefficient $M \geqq 1$ on $R^{*}$, and suppose

$$
\left|\zeta^{*}(x)\right| \leqq k, x \in R^{*}
$$

Then there is a function $\zeta$ on $\bar{R}$ which satisfies a Lipschitz condition with coefficient $M$ on $\bar{R}$, coincides with $\zeta^{*}$ on $R^{*}$, and is zero except on a set of measure at most

$$
m(R) \cdot\left[1-\left(1-h^{-1} k\right)^{\nu}\right] .
$$

Proof. Let $R_{1}$ be the cell with center at $x_{0}$ and edges $2\left(h^{\alpha}-k\right), \alpha=1$, $\cdots, \nu$. Then, since $h=\min h^{\alpha}$, we have

$$
m\left(R_{1}\right) \geqq m(R) \cdot\left(1-h^{-1} k\right)^{\nu}
$$

Define $\zeta_{1}=0$ on $\bar{R}_{1}$ and equal to $\zeta^{*}$ on $R^{*}$. Then

$$
\left|\zeta_{1}\left(x_{1}\right)-\zeta_{1}\left(x_{2}\right)\right| \leqq\left|x_{1}-x_{2}\right| \quad \text { if } x_{1} \in \bar{R}_{1}, x_{2} \in R^{*}
$$

Thus $\zeta_{1}$ satisfies a uniform Lipschitz condition with coefficient $M$ on $\bar{R}_{1} \cup R^{*}$. By a well known theorem, there exists an extension of $\zeta_{1}$ to $\bar{R}$ (the whole space in fact) which satisfies the same Lipschitz condition.

Lemma 2.2. Suppose the vectors $\zeta_{n} \rightarrow 0$ (in our sense) on $\bar{R}$ and suppose $f$ is quasi-convex in $p$. Then if $p_{0}$ is a constant vector we have

$$
m(R) f\left(p_{0}\right) \leqq \liminf _{n \rightarrow \infty} \int_{R} f\left[p_{0}+\pi_{n}(x)\right] d x
$$

Proof. For all sufficiently large $n$, we have $k_{n}<h$, and $k_{n} \rightarrow 0, k_{n}$ being the maximum of $\left|\zeta_{n}(x)\right|$ for $x \in R^{*}$. For each $n$ for which $k_{n}<h$, let $\eta_{n}$ be the
function of the preceding lemma which coincides on $R^{*}$ with $\zeta_{n}$, and let $\omega_{n}$ $=\zeta_{n}-\eta_{n}$. Then if each $\zeta_{n}$ satisfies a uniform Lipschitz condition with coefficient $M \geqq 1$ on $\bar{R}$, then each $\eta_{n}$ and $\omega_{n}$ satisfies one with coefficient $M$ and $2 M$, respectively. Moreover, each derivative $\eta_{n x^{a}}^{i}$ is uniformly bounded and $\eta_{n x}^{i} \longrightarrow 0$ almost everywhere. Since $f$ is uniformly continuous on any bounded portion of $p$-space, we see that

$$
\lim _{n \rightarrow \infty} \int_{R}\left|f\left(p_{0 \alpha}^{i}+\eta_{n x \alpha}^{i}+\omega_{n x \alpha}^{i}\right)-f\left(p_{0 \alpha}^{i}+\omega_{n x \alpha}^{i}\right)\right| d x=0
$$

But the result then follows, since, for each $n$, we have

$$
\int_{R} f\left(p_{0 \alpha}^{i}+\omega_{n x \alpha}^{i}\right) \mathrm{d} \mathbf{x} \geqq m(R) f\left(p_{0}\right)
$$

because of the quasi-convexity of the function $f$.
Theorem 2.2. Suppose $f$ is continuous in $(x, z, p)$ for all $(x, z, p)$ and is quasi-convex in $p$ for each $(x, z)$. Suppose also that $z_{n} \rightarrow z_{0}$ on the bounded domain D. Then

$$
I\left(z_{0}, D\right) \leqq \liminf _{n \rightarrow \infty} I\left(z_{n}, D\right)
$$

Proof. Let $\epsilon$ be any positive number. For each positive integer $k$, let $D_{k}$ consist of all the hypercubes of edge $2^{-k}$ whose faces lie along hyperplanes $x^{a}=2^{-k} i^{\alpha}$ (each $i^{\alpha}$ an integer) which lie in $D$. Since all the points $\left[x, z_{0}(x)\right.$, $\left.p_{0}(x)\right]$ and $\left[x, z_{n}(x), p_{n}(x)\right]$ for $x \in D$ lie in a bounded portion of $(x, z, p)$ space, we may choose $k_{1}$ so large that

$$
\begin{equation*}
\int_{D-D_{k_{1}}}\left|f\left(x, z_{n}, p_{n}\right)\right| d x<\epsilon / 5, \int_{D-D_{k_{1}}}\left|f\left(x, z_{0}, p_{0}\right)\right| d x<\epsilon / 5 \tag{2.1}
\end{equation*}
$$

for all $n$.
Let the hypercubes of $D_{k_{1}}$ be $R_{1}, \cdots, R_{N}$. For each $k \geq k_{1}$, let $R_{k i}$, $i=1, \cdots, N \cdot 2^{\nu\left(k-k_{1}\right)}$, be all the hypercubes of side $2^{-k}$ described above which lie in $D_{k_{1}}$. For each such $k$, define $x_{k}^{*}(x), z_{k}^{*}(x), p_{k}^{*}(x)$ on $D_{k_{1}}$ by

$$
\begin{aligned}
& x_{k}^{*}(x)=\left[m\left(R_{k i}\right)\right]^{-1} \int_{R_{k i}} x d x, z_{k}^{*}(x)=\left[m\left(R_{k i}\right)\right]^{-1} \int_{R_{k i}} z_{0}(x) d x \\
& p_{k}^{*}(x)=\left[m\left(R_{k i}\right)\right]^{-1} \int_{R_{k i}} p_{0}(x) d x \\
& r_{k}(x)=\left\{\left|x_{k}^{*}(x)-x\right|^{2}+\left|z_{k}^{*}(x)-z_{0}(x)\right|^{2}+\left|p_{k}^{*}(x)-p_{0}(x)\right|^{2}\right\}^{1 / 2}
\end{aligned}
$$

where $\quad x \in R_{k i}$. Let $\zeta_{n}(x)=z_{n}(x)-z_{0}(x), \pi_{n}(x)=p_{n}(x)-p_{0}(x)$. Then, on $D_{k_{1}}$,

$$
\begin{aligned}
& f\left[x, z_{n}(x), p_{n}(x)\right]-f\left[x, z_{0}(x), p_{0}(x)\right] \\
= & \left\{f\left[x, z_{n}(x), p_{n}(x)\right]-f\left[x, z_{0}(x), p_{n}(x)\right]\right\} \\
+ & \left\{f\left[x, z_{0}(x), p_{0}(x)+\pi_{n}(x)\right]-f\left[x_{k}^{*}(x), z_{k}^{*}(x), p_{k}^{*}(x)+\pi_{n}(x)\right]\right\} \\
- & \left\{f\left[x, z_{0}(x), p_{0}(x)\right]-f\left[x_{k}^{*}(x), z_{k}^{*}(x), p_{k}^{*}(x)\right]\right\} \\
+ & \left\{f\left[x_{k}^{*}(x), z_{k}^{*}(x), p_{k}^{*}(x)+\pi_{n}(x)\right]-f\left[x_{k}^{*}(x), p_{k}^{*}(x), p_{k}^{*}(x)\right]\right\} .
\end{aligned}
$$

Now, all the arguments of $f$ occurring in (2.3) for $x \in D_{k_{1}}$ lie in a bounded closed cell in ( $x, z, p$ )-space over which $f$ is uniformly continuous. Let

$$
\epsilon(\rho)=\max \left|f\left(x^{\prime}, z^{\prime}, p^{\prime}\right)-f\left(x^{\prime \prime}, z^{\prime \prime}, p^{\prime \prime}\right)\right|, \quad \rho \geqq 0
$$

for all $\left(x^{\prime}, z^{\prime}, p^{\prime}\right)$ and ( $\left.x^{\prime \prime}, z^{\prime \prime}, p^{\prime \prime}\right)$ in this cell with

$$
\left|x^{\prime}-x^{\prime \prime}\right|^{2}+\left|z^{\prime}-z^{\prime \prime}\right|^{2}+\left|p^{\prime}-p^{\prime \prime}\right|^{2} \leqq \rho^{2} .
$$

then $\epsilon(\rho)$ is continuous for $\rho \geqq 0$ with $\epsilon(0)=0$. Then, for each $n$ and each $k \geqq k_{1}$, we have
$\left|f\left[x, z_{n}(x), p_{n}(x)\right]-f\left[x, z_{0}(x), p_{n}(x)\right]\right| \leqq \epsilon\left(\left|z_{n}(x)-z_{0}(x)\right|\right)$,
$\left|f\left[x, z_{0}(x), p_{0}(x)+\pi_{n}(x)\right]-f\left[x_{k}^{*}(x), z_{k}^{*}(x), p_{k}^{*}(x)+\pi_{n}(x)\right]\right| \leqq \epsilon\left[r_{k}(x)\right]$,
$\left|f\left[x, z_{0}(x), p_{0}(x)\right]-f\left[x_{k}^{*}(x), z_{k}^{*}(x), p_{k}^{*}(x)\right]\right| \leqq \epsilon\left[r_{k}(x)\right]$.
Now, the $r_{k}(x)$ are uniformly bounded on $D_{k_{1}}$ and tend to zero almost everywhere on $D_{k_{1}}$. Hence we may choose a $k \geqq k_{1}$ so large that
$\int_{D_{k_{1}}}\left|f\left[x, z_{0}(x), p_{0}(x)+\pi_{n}(x)\right]-f\left[x_{k}^{*}(x), z_{k}^{*}(x), p_{k}^{*}(x)+\pi_{n}(x)\right]\right| d x<\epsilon / 5$,
(2.4)
$\int_{D_{k_{1}}}\left|f\left[x, z_{0}(x), p_{0}(x)\right]-f\left[x_{k}^{*}(x), z_{k}^{*}(x), p_{k}^{*}(x)\right]\right| d x<\epsilon / 5$,
for all $n$. Since $z_{n}$ converges uniformly to $z_{0}$, there is an $n_{1}$ such that

$$
\begin{equation*}
\int_{D_{k_{1}}}\left|f\left[x, z_{n}(x), p_{n}(x)\right]-f\left[x, z_{0}(x), p_{n}(x)\right]\right| d x<\epsilon / 5, \quad n>n_{1} \tag{2.5}
\end{equation*}
$$

Finally, since $x_{k}^{*}(x)$, and so on, are constant on each $R_{k i}$, and $f$ is quasi-convex, we conclude from the previous lemma that

$$
\liminf _{n \rightarrow \infty} \int_{D_{k_{1}}}\left\{f\left[x_{k}^{*}(x), z_{k}^{*}(x), p_{k}^{*}(x)+\pi_{n}(x)\right]-f\left[x_{k}^{*}(x), z_{k}^{*}(x), p_{k}^{*}(x)\right]\right\} d x \geqq 0
$$

Using (2.3)-(2.5) and the above inequality, we see that

$$
\liminf _{n \rightarrow \infty} I\left(z_{n}, D\right) \geqq I\left(z_{0}, D\right)-\epsilon .
$$

Since $\epsilon$ is any positive number, the result follows.
3. Lower semicontinuity and weak convergence in $\Re_{s}(s \geqq 1)$. In this section, we discuss additional conditions which with the quasi-convexity of $f$ in $p$ are sufficient to guarantee the lower semicontinuity of $l(z, D)$ with respect to weak convergence in $\mathbb{F}_{s}$ on $D$.

Definition 3.1. Suppose $\zeta$ is of class $\Re_{s}$ on the bounded domain $D$ and suppose $R$ is a cell with $\bar{R} \subset D$. Then $\zeta$ is said to be strongly of class $\Re_{s}$ on $R^{*}$ if and only if $\bar{\zeta}$ is of class $\Re_{s}$ in $x_{\alpha}^{\prime}$ on each face $x^{\alpha}=$ const. of $R^{*}$ and there is a sequence $\zeta_{n}$ of class $C^{\prime}$ on $\bar{R}$ such that

$$
\bar{D}_{s}\left(\zeta_{n}-\zeta, R\right) \rightarrow 0, \bar{D}_{s}\left(\zeta_{n}-\bar{\zeta}, R^{*}\right) \rightarrow 0
$$

Lemma 3.1. Suppose $\zeta$ is of class $\Re_{s}(s \geqq 1)$ on the bounded domain $D$. For each $\alpha, 1 \leqq \alpha \leqq \nu$, let $\left(a^{\alpha}, b^{\alpha}\right)$ be the open interval projection of $D$ on the $x^{\alpha}$ axis. Then there exist sets $Z^{\alpha}$ of measure zero such that if $R: c^{\alpha} \leqq x^{\alpha} \leqq d^{\alpha}$ $(\alpha=1, \cdots, \nu)$ is any closed cell in $D$ with

$$
c^{\alpha} \in\left(a^{a}, b^{\alpha}\right)-Z^{a}, \quad d^{a} \in\left(a^{\alpha}, b^{\alpha}\right)-Z^{\alpha} \quad(\alpha=1, \cdots, \nu)
$$

then $\zeta$ is strongly of class $\Re_{s}$ on $R^{*}$.
Proof. Let $R^{\prime}$ be any rational cell in $D$ (that is, $R=[C, D]$ with $C^{a}, D^{a}$ rational). In [1], Lemma 5.1, we have seen that if $\zeta$ is of class $\Re_{s}$ on $D$, then

$$
\begin{equation*}
\lim _{h \rightarrow 0} \bar{D}_{s}\left(\zeta_{h}-\zeta, R\right)=0 \tag{3.1}
\end{equation*}
$$

For each $\alpha$, define

$$
\phi_{h}^{\alpha}\left(x^{\alpha}, R^{\prime}\right)=\int_{C_{\alpha}^{\prime}}^{D_{\alpha}^{\prime}}\left\{\left|\zeta_{h}-\bar{\zeta}\right|^{s}+\left[\sum_{\substack{\beta=1 \\ \beta \neq \alpha}}^{\nu}\left|\zeta_{h x} \beta-\bar{\zeta}_{x}\right|^{2}\right]^{s / 2}\right\} d x_{a}^{\prime}
$$

Since $\bar{\zeta}$ is obviously of class $\Re_{s}^{\prime}$ in $x_{\alpha}^{\prime}$ for almost all $x^{\alpha}$ on $\left[C^{\alpha}, D^{\alpha}\right], \phi_{h}^{\alpha}\left(x^{\alpha}, R^{\prime}\right)$ is defined for almost all $x^{\alpha}$ and

$$
\lim _{h \rightarrow 0} \int_{C^{\alpha}}^{D^{\alpha}}\left|\phi_{h}^{\alpha}\left(x^{\alpha}, R^{\prime}\right)\right| d x^{\alpha}=0
$$

By arranging the rational cells $R^{\prime}$ in some order and choosing successive subsequences, we may choose (on account of (3.1)) a final sequence $h_{n} \rightarrow 0$ such that $\phi_{h_{n}}^{a}\left(x^{\alpha}, R^{\prime}\right) \rightarrow 0$ and $\bar{\zeta}$ is of class $\Re_{s}^{\prime}$ in $x_{\alpha}^{\prime}$ on $\left[C_{a}^{\prime}, D_{\alpha}^{\prime}\right]$ for each $x^{\alpha}$ not in a set $Z^{\alpha}\left(R^{\prime}\right)$ of measure zero $\quad(\alpha=1, \cdots, \nu)$. Now let

$$
Z^{\alpha}=U Z^{\alpha}\left(R^{\prime}\right)
$$

then

$$
m\left(Z^{\alpha}\right)=0 \quad(\alpha=1, \cdots, \nu)
$$

Now suppose $R$ is one of the cells described in the lemma. Then it lies in some rational cell $R^{\prime}$ and we may take $\zeta_{n}=\zeta_{h_{n}}$.

Lemma 3.2. Suppose $R$ is a cell with edges $\left(2 h^{1}\right), \cdots,\left(2 h^{\nu}\right)$ and center $x_{0}$. Let

$$
h=\min _{1 \leqq}^{1 \leqq} \quad h^{\alpha}, \quad K=h^{-1}\left(h^{\alpha} h^{\alpha}\right)^{1 / 2} .
$$

Suppose also that $0<k<h$, that $\zeta^{*}$ is of class $\Re_{s}$ on an open domain containing $\bar{R}$ in its interior, and that $\zeta^{*}$ is strongly of class $\Re_{s}$ on $R^{*}$ with

$$
\int_{R^{*}}\left|\zeta^{*}\right|^{s} d S \leqq k^{s}, \quad D_{s}\left(\zeta^{*}, R^{*}\right) \leqq M^{s} \quad(s \geqq 1)
$$

Then there is a function $\zeta$ of class $\beta_{s}$ on $R$ which coincides with $\zeta^{*}$ on $R^{*}$, is zero except on a set of measure

$$
m(R) \cdot\left[1-\left(1-h^{-1} k\right)^{\nu}\right],
$$

and satisfies

$$
D_{s}(\zeta, R) \leqq \tau_{s} h^{-1} k\left(1+K^{s} M^{s}\right), \quad \tau_{s}= \begin{cases}2^{s / 2} & (s \leqq 2) \\ 2^{s-1} & (s \leqq 2)\end{cases}
$$

Proof. For each $x \in \bar{R}, x \neq x_{0}$, let $x^{*}(x)$ be the intersection of the ray $\overrightarrow{x_{0} x}$ with $R^{*}$, and for each $x \in \bar{R}$ define

$$
r(x)=\left\{\begin{array}{cl}
0 & \left(x=x_{0}\right) \\
\left|x^{*}(x)-x_{0}\right|^{-1} \cdot\left|x-x_{0}\right| & \left(x \neq x_{0}\right)
\end{array}\right.
$$

Let $\Pi_{\alpha}^{ \pm}$be the pyramid in $\bar{R}$ with vertex $x_{0}$ and base the face $F_{\alpha}^{ \pm}$where

$$
x^{\alpha}=x_{0}^{\alpha} \pm h^{\alpha} .
$$

On the pyramid $\Pi_{\nu}^{+}$, introduce coordinates $\xi^{1}, \cdots, \xi^{\nu-1}, r$ by

$$
x^{\nu}=x_{0}^{\nu}+r h^{\nu}, x^{\gamma}=x_{0}^{\gamma}+r \xi^{\gamma} \quad(0 \leqq r \leqq 1, \gamma=1, \cdots, \nu-1)
$$

Then, if $r$ and $\xi^{\gamma}$ are considered as functions of $x$, we have

$$
r(x)=r, x^{*}(x)=\left[\xi^{1}(x)+x_{0}^{1}, \cdots, \xi^{\nu-1}(x)+x_{0}^{\nu-1}, h^{\nu}+x_{0}^{\nu}\right]
$$

Similar coordinate systems may be set up on each of the other $\Pi_{\alpha}^{ \pm}$.
Define

$$
\phi(r)= \begin{cases}0 & \left(0 \leqq r \leqq 1-k h^{-1}\right) \\ h k^{-1}\left(r-1+k h^{-1}\right) & \left(1-k h^{-1} \leqq r \leqq 1\right)\end{cases}
$$

Choose a sequence $\zeta_{n}^{*}$ satisfying the conditions of Definition 3.1; and for each $n$, define

$$
\zeta_{n}(x)=\phi[r(x)] \cdot \zeta_{n}^{*}\left[x^{*}(x)\right]
$$

Then each $\zeta_{n}(x)$ is of class $D^{\prime}$ on $\bar{R}$.
We now compute the derivatives of $\zeta_{n}$ on each pyramid $\Pi_{a}^{ \pm}$taking $\Pi_{\nu}^{+}$as an example. Then

$$
\begin{array}{ll}
\zeta_{n x} \gamma & =r^{-1} \phi(r) \zeta_{n \xi}^{*} \\
\zeta_{n x \nu}=\left(h^{\nu}\right)^{-1} \phi^{\prime}(r) \zeta_{n}^{*}-\left(h^{\nu}\right)^{-1} r \phi(r) \xi^{\gamma} \zeta_{n \xi^{\gamma}}^{*}(\gamma \text { summed from } 1 \text { to } \nu-1) .
\end{array}
$$

Then, since $r^{-1} \phi(r) \leqq 1$ and $\phi^{\prime}(r)=k^{-1} h$ for $1-h k^{-1} \leqq r \leqq 1$,

$$
\begin{aligned}
\left|\pi_{n}(x)\right|^{2} & \leqq\left(\zeta_{n \xi}^{* i} \gamma \zeta_{n \xi}^{* i}\right)+2 k^{-2}\left|\zeta_{n}^{*}\right|^{2}+2\left(h^{\nu}\right)^{-2}\left(\xi^{\gamma} \xi^{\gamma}\right)\left(\zeta_{n \xi}^{* i} \gamma \zeta_{n \xi}^{* i}\right) \\
& \leqq 2\left[k^{-2}\left|\zeta_{n}^{*}\right|^{2}+K^{2}\left(\zeta_{n \xi}^{* i} \gamma \zeta_{n \xi}^{* i}\right)\right] \quad \quad \text { (n not summed) }
\end{aligned}
$$

Using the inequality

$$
\left(a^{2}+b^{2}\right)^{s / 2} \leqq \sigma_{s}\left(|a|^{s}+|b|^{s}\right), \quad \sigma_{s} \leqq \begin{cases}1 & (s \leqq 2) \\ 2^{(s-2) / 2} & (s \leqq 2)\end{cases}
$$

we obtain

$$
\begin{aligned}
D_{s}\left(\zeta_{n}, \Pi_{\nu}^{+}\right) \leqq & \tau_{s} \int_{1-k h^{-1}}^{1} r^{\nu-1} d r \int_{F_{\nu}^{+}}\left[k^{-s}\left|\zeta_{n}^{*}\right|^{s}\right. \\
& \left.\left.+K^{s}\left(\zeta_{n \xi^{\gamma}}^{* i} \zeta_{n \xi}^{* i}\right)\right)^{s / 2}\right] d s \\
\leqq & \tau_{s} h^{-1} k\left[k^{-s} \int_{F_{\nu}^{+}}\left|\zeta_{n}^{*}\right|^{s} d s+K^{s} D_{s}\left(\zeta_{n}^{*}, F_{\nu}^{+}\right)\right]
\end{aligned}
$$

Also

$$
\begin{aligned}
\int_{\Pi_{\nu}^{+}}\left|\zeta_{n}\right|^{s} d x & =\int_{1-k h^{-1}}^{1} r^{\nu-1} \phi^{s}(r) d r \int_{F_{\nu}^{+}}\left|\zeta_{n}^{*}\right|^{s} d S \\
& \leqq h^{-1} k \int_{F_{\nu}^{+}}\left|\zeta_{n}^{*}\right|^{s} d S
\end{aligned}
$$

Adding these results for all the $\Pi_{\alpha}^{ \pm}$, we obtain the result for each $n$; and also $\bar{D}_{s}\left(\zeta_{n}, R\right)$ is uniformly bounded. Thus, we may extract a subsequence which tends weakly in $\Re_{s}$ to some function $\zeta$ of class $\Re_{s}$ on $R$. Since each $\zeta_{n}=\zeta_{n}^{*}$ on $R^{*}, \zeta_{n}^{*}$ tends strongly in $L_{s}$ to $\bar{\zeta}^{*}$ on $R^{*}$, we see from [2], Theorem 8.5, that $\zeta=\bar{\zeta}^{*}$ on $R^{*}$. From the lower semicontinuity of $D_{s}$ (see [2], Theorem 8.2), the result follows.

Lemma 3.3. Suppose $f$ is quasi-convex and of class $C^{\prime}$ for all $p$, and suppose for all $p$ that

$$
\sum_{i, \alpha}\left(\begin{array}{c}
\left.f_{p_{\alpha}^{i}}\right)^{2} \leqq K^{2}\left(|p|^{s-1}+1\right)^{2}
\end{array}(s \geqq 1)\right.
$$

If $p_{0}$ is any constant vector, $D$ is any bounded domain, and $\zeta$ is of class $\Re_{s}$ on $D$ and vanishes on $D^{*}$, then $f\left[p_{0}+\pi(x)\right]$ is summable over $D$ and

$$
\int_{D} f\left[p_{0}+\pi(x)\right] d x \geqq m(D) \cdot f\left(p_{0}\right)
$$

Proof. There exists a sequence of functions $\zeta_{n}$, each of class $C^{\prime}$ on $D$ and vanishing on and near $D^{*}$, such that $\bar{D}_{s}\left(\zeta_{n}-\zeta, D\right) \longrightarrow 0$ (see [2], Definition 9.1). For each $n$ and almost all $x$ on $D$, we have

$$
\begin{aligned}
& \left|f\left[p_{0}+\pi_{n}(x)\right]-f\left[p_{0}+\pi(x)\right]\right|= \\
& \qquad\left|\left[\pi_{n \alpha}^{i}(x)-\pi_{\alpha}^{i}(x)\right] \int_{0}^{1} f_{p_{\alpha}^{i}}\left[p_{0}+(1-t) \pi(x)+t_{n}(x)\right] d t\right|
\end{aligned}
$$

$$
\leqq\left|\pi_{n}(x)-\pi(x)\right| \cdot K \cdot \int_{0}^{1}\left\{\left|(1-t) p_{0}+\pi(x)+t p_{0}+\pi_{n}(x)\right|-+1\right\} d t
$$

$$
\stackrel{(3.2)}{\leqq} K\left|\pi_{n}(x)-\pi(x)\right|\left\{h_{s}\left|p_{0}+\pi(x)\right|^{s-1}+h_{s}\left|p_{0}+\pi_{n}(x)\right|^{s-1}+1\right\}
$$

where

$$
h_{s}= \begin{cases}s^{-1} & (1 \leqq s \leqq 2) \\ s^{-1} 2^{s-2} & (s \leqq 2)\end{cases}
$$

Using the Hölder inequality, and so on, and the strong convergence in $\Re_{s}$, we see that

$$
\lim _{n \rightarrow \infty} \int_{D} f\left[p_{0}+\pi_{n}(x)\right] d x=\int_{D} f\left[p_{0}+\pi(x)\right] d x
$$

Since $f$ is quasi-convex, the result follows.
Lemma 3.4. Suppose that fatisfies the hypotheses of Lemma 3.3. Suppose also that each $\zeta_{n}$ is of class $\Re_{s}$ on a domain $D$ and is strongly of class $\Re_{s}$ on $R^{*}, \bar{R} \subset D$, with
$\lim _{n \rightarrow \infty} \int_{R^{*}}\left|\zeta_{n}\right|^{s} d S=0, D_{s}\left(\zeta_{n}, R^{*}\right) \leqq M^{s}, D_{s}\left(\zeta_{n}, R\right) \leqq M^{s} \quad(n=1,2 \ldots)$.
Then for each $p_{0}, f\left[p_{0}+\pi_{n}(x)\right]$ is summable for all sufficiently large $n$, and

$$
\liminf _{n \rightarrow \infty} \int_{R} f\left[p_{0}+\pi_{n}(x)\right] d x \geqq m(R) \cdot f\left(p_{0}\right), \pi_{n \alpha}^{i}(x)=\zeta_{n x^{\alpha}}^{i}(x)
$$

Proof. For each n, let

$$
k_{n}=\left[\int_{R^{*}}\left|\zeta_{n}\right|^{s} d S\right]^{1 / s},
$$

and let $K$ and $h$ be the quantities of Lemma 3.2 for $R$. Since $k_{n} \longrightarrow 0$, we have $k_{n}<h$ for all $n>$ some $n_{1}$. For each such $n$, let $\eta_{n}$ be the function of Lemma 3.2 which coincides on $R^{*}$ with $\zeta_{n}$, and let

$$
\chi_{n}=\zeta_{n}-\eta_{n}, \kappa_{n \alpha}^{i}=\eta_{n x^{a}}^{i}, \omega_{n \alpha}^{i}=\chi_{n x^{\alpha}}^{i}
$$

Then, since $\chi_{n}=0$ on $R^{*}$, we have

$$
\int_{R} f\left[\pi_{0}+\omega_{n}(x)\right] d x \geqq m(R) f\left(p_{0}\right)
$$

As in (3.2), we see that, for each $n$, and almost all $x$ on $D$,

$$
\begin{aligned}
& \left|f\left[p_{0}+\omega_{n}(x)+\kappa_{n}(x)\right]-f\left[p_{0}+\omega_{n}(x)\right]\right| \\
& \quad \leqq K \cdot\left|\kappa_{n}(x)\right| \cdot\left(h_{s}\left|p_{0}+\omega_{n}(x)+\kappa_{n}(x)\right|^{s-1}+h_{s}\left|p_{0}+\omega_{n}(x)\right|^{s-1}+1\right) \\
& \quad \leqq K \cdot\left|\kappa_{n}(x)\right| \cdot\left[\left(1+s h_{s}\right) h_{s}\left|p_{0}+\pi_{n}(x)\right|^{s-1}+s h_{s}^{2}\left|\kappa_{n}(x)\right|^{s-1}+1\right] .
\end{aligned}
$$

Using the Hölder inequality, and so on, we see that

$$
\lim _{n \rightarrow \infty} \int_{R}\left|f\left[p_{0}+\pi_{n}(x)\right]-f\left[p_{0}+\omega_{n}(x)\right]\right| d x=0
$$

from which the result follows.
Theorem 3.1. Suppose $f$ is of class $C^{\prime}$ in $(x, z, p)$ and quasi-convex in $p$. Suppose also that there are numbers $k$ and $K, K>0$, such that
(i) $f(x, z, p) \geqq k$,
(iii) $f_{x^{\alpha}} f_{x^{\alpha}} \leqq K^{2}\left(|p|^{s}+1\right)^{2}$
(ii) $f_{p_{\alpha}^{i}} f_{p_{\alpha}^{i}} \leqq K^{2}\left(|p|^{s-1}+1\right)^{2}$,
(iv) $f_{z^{i}} f_{z^{i}} \leqq K^{2}\left(|p|^{s}+1\right)^{2}$.
for all ( $x, z, p$ ).

Suppose also that $z_{n} \rightarrow z_{0}$ weakly in $\Re_{s}$ on the bounded domain $D$ and that either
(a) each $z_{n}$ and $z_{0}$ are continuous on $D$ and $z_{n}$ converges uniformly to $z_{0}$ on each closed set interior to $D$, or
(b) the set functions $D_{s}\left(z_{n}, e\right)$ are uniformly absolutely continuous on each closed set interior to $D$.

Then

$$
I\left(z_{0}, D\right) \leqq \liminf _{n \rightarrow \infty} I\left(z_{n}, D\right)
$$

REMARK. If $s=1$, weak convergence in $\Re_{s}$ implies the hypothesis (b). Proof. We note first that hypothesis (ii) implies

$$
\begin{align*}
& |f(x, z, p)-f(x, z, 0)|=\left|p_{\alpha}^{i} \int_{0}^{1} f_{p_{\alpha}^{i}}\left(x, z, t p_{\alpha}^{i}\right) d t\right|  \tag{3.3}\\
& \quad \leqq|p| \cdot \int_{0}^{1}\left[f_{p_{\alpha}^{i}} f_{p_{\alpha}^{i}}\left(x, z, t p_{\alpha}^{i}\right)\right]^{1 / 2} d t
\end{align*}
$$

$$
\leqq|p| \cdot \int_{0}^{1} K\left(t^{s-1} p^{s-1}+1\right) d t \leqq K\left(s^{-1}|p|^{s}+|p|\right)
$$

Also, hypotheses (iii) and (iv) similarly imply

$$
\begin{equation*}
|f(x, z, 0)-f(0,0,0)| \leqq K(|x|+|z|) \tag{3.4}
\end{equation*}
$$

Thus, for all $(x, z, p)$, we have

$$
\begin{equation*}
|f(x, z, p)| \leqq|f(0,0,0)|+K\left(|x|+|z|+s^{-1}|p|^{s}+|p|\right) \tag{3.5}
\end{equation*}
$$

Therefore $I\left(z_{0}, D\right)$ and the $I\left(z_{n}, D\right)$ are uniformly bounded.
For each $\alpha(1 \leqq \alpha \leqq \nu)$, let $\left(a^{\alpha}, b^{\alpha}\right)$ be the open interval projection of $D$ on the $x^{\alpha}$ axis and let $Z_{0}^{\alpha}$ and $Z_{n}^{\alpha}$ be the sets of Lemma 3.1 for $z_{0}$ and $z_{n}$. Also for each $\alpha, n, k$, let $E_{n, k}^{\alpha}$ be the set of $x^{\alpha}$ in $\left(a^{a}, b^{a}\right)-Z_{n}^{\alpha}$, where

$$
\bar{D}_{s}\left(\bar{z}_{n}, D_{x^{\alpha}}\right) \leqq k
$$

$D_{x \alpha}$ being the set of $x_{\alpha}^{\prime}$ such that $\left(x_{\alpha}^{\prime}, x^{\alpha}\right) \in D$. Suppose that $\bar{D}_{s}\left(z_{n}, D\right) \leqq M$, some uniform bound existing because of the weak convergence. Let

$$
Z_{n, k}^{a}=\left(a^{\alpha}, b^{\alpha}\right)-E_{n, k}^{a}
$$

Then

$$
m\left(Z_{n, k}^{a}\right) \leqq M k^{-1}, m\left(E_{n, k}^{\alpha}\right) \geqq\left(b^{\alpha}-a^{\alpha}\right)-M k^{-1}
$$

For each $\alpha$, let

$$
E^{a}=\bigcup_{k=1}^{\infty} \bigcap_{N=1}^{\infty} \bigcup_{n=N}^{\infty} E_{n, k}^{a}, \tilde{Z}_{0}^{a}=\left(a^{\alpha}, b^{\alpha}\right)-E^{\alpha} \cup Z_{0}^{a} \cup \bigcup_{n=1}^{\infty} Z_{n}^{a} .
$$

Then $m\left(Z_{0}^{a}\right)=0$. For each $\alpha$, each natural number $n$, and each integer $i$, define $Z_{n, i}^{\alpha}$ as the set of all $x^{a}$ such that $x^{\alpha}-i \cdot 2^{-n} \in \tilde{Z}_{0}^{\alpha}$, and define

$$
Z^{a}=\bigcup_{n, i} Z_{n, i}^{a}
$$

Then $m\left(Z^{\alpha}\right)=0$.
Now, choose a point $x_{0}$ such that $x_{0}^{\alpha}$ is not in $Z^{\alpha}(\alpha=1, \cdots, \nu)$. For each natural number $k$, let $C_{k}$ be the totality of hypercubes of side $2^{-k}$ bounded by hyperplanes of the form $x^{\alpha}=x_{0}^{\alpha}+i \cdot 2^{-k}$. None of the numbers $x_{0}^{\alpha}+i \cdot 2^{-k}$ is in $Z^{\alpha}$ and, moreover, $\bar{z}_{0}$ and each $\bar{z}_{n}$ is strongly of class $\wp_{s}$ on $R^{*}$ with
$\bar{D}_{s}\left(\bar{z}_{n}, R\right)$ uniformly bounded for infinitely many values of $n, R$ being any hypercube of any $C_{f_{k}}$. Since the totality of these hypercubes is countable, we may choose a subsequence, still called $z_{n}$, such that $l\left(z_{n}, D\right)$ tends to the former lim inf, $\overline{z_{n}} \longrightarrow \bar{z}_{0}$ almost everywhere on $D$, and $\bar{D}_{s}\left(z_{n}, R^{*}\right)$ is uniformly bounded for $R$ of any $G_{k}$ in $D$. Since $z_{n} \rightarrow z$ in $S_{s}$, we also have

$$
\lim _{n \rightarrow \infty} \int_{R^{*}}\left|z_{n}-z_{0}\right|^{s} d S=0
$$

for each such $R$.
Now, we first consider the alternative (a). Let $\epsilon$ be any positive number. For each $k$, let $D_{k}$ be the union of all the cells of $\delta_{k}$ which are interior to $D$. Since $f$ is bounded below and $I\left(z_{0}, D\right)$ is finite, we first choose $k_{1}$ so large that

$$
I\left(z_{n}, D-D_{k_{1}}\right)>-\epsilon / 5 \quad(n=1,2, \cdots)
$$

$$
\begin{equation*}
l\left(z_{0}, D_{k_{1}}\right)>I\left(z_{0}, D\right)-\epsilon / 5 \tag{3.6}
\end{equation*}
$$

For this $k_{1}$, let $R_{1}, \ldots, R_{q}$ be the cells of $D_{k_{1}}$ and for each $k \geqq k_{1}$, let

$$
R_{k i} \quad\left(i=1, \cdots, q \cdot 2^{\nu\left(k-k_{1}\right)}\right)
$$

be the cells of $\mathscr{f}_{k}$ in $D_{k_{1}}$ For each $k$, define $x_{k}^{*}(x), z_{k}^{*}(x)$, and $p_{k}^{*}(x)$ on $D_{k_{1}}$ by (2.9). Then, from (ii), (iii), and (iv), it follows that

$$
\begin{aligned}
& \left|f\left[x, z_{0}(x), p_{0}(x)\right]-f\left[x_{k}^{*}(x), z_{k}^{*}(x), p_{k}^{*}(x)\right]\right| \\
& \quad \leqq K\left(\left|p_{0}(x)\right|^{s}+1\right) \cdot\left(\left|x-x_{k}^{*}(x)\right|+\left|z_{0}(x)-z_{k}^{*}(x)\right|\right) \\
& \quad
\end{aligned}
$$

where

$$
h_{s}= \begin{cases}s^{-1} & (1 \leqq s \leqq 2) \\ s^{-1} \cdot 2^{s-2} & (s \leqq 2)\end{cases}
$$

the method of proof is similar to that of (3.3). If we let

$$
\zeta_{n}=z_{n}-z_{0}, \pi_{n}=\dot{p}_{n}-p_{0},
$$

we see similarly that

$$
\left|f\left[x, z_{0}(x), p_{0}(x)+\pi_{n}(x)\right]-f\left[x_{k}^{*}(x), z_{k}^{*}(x), p_{k}^{*}(x)+\pi_{n}(x)\right]\right|
$$

$$
\begin{align*}
& \leqq K\left(\left|p_{n}(x)\right|^{s}+1\right) \quad\left(\left|x-x_{k}^{*}(x)\right|+\left|z_{0}(x)-z_{k}^{*}(x)\right|\right)  \tag{3.8}\\
& +K\left(h_{s}\left|p_{n}(x)\right|^{s-1}+h_{s}\left|p_{k}^{*}(x)+\pi_{n}(x)\right|^{s-1}+1\right) \cdot\left|p_{0}(x)-p_{k}^{*}(x)\right|
\end{align*}
$$

$$
\begin{align*}
& \left|f\left[x, z_{n}(x), p_{n}(x)\right]-f\left[x, z_{0}(x), p_{n}(x)\right]\right|  \tag{3.9}\\
& \quad \leqq K\left(\left|p_{n}(x)\right|^{s}+1\right) \cdot\left|z_{n}(x)-z_{0}(x)\right| .
\end{align*}
$$

Now, by the Hölder inequality on each $R_{k i}$, we see that

$$
\begin{equation*}
\int_{D_{k_{1}}}\left|p_{k}^{*}(x)\right|^{s} d x \leqq \int_{D_{k_{1}}}\left|p_{0}(x)\right|^{s} d x . \tag{3.10}
\end{equation*}
$$

By applying the Minkowski inequality, we see that the integrals

$$
\begin{equation*}
\int_{D_{k_{1}}}\left|\pi_{n}(x)\right|^{s} d x, \int_{D_{k_{1}}}\left|p_{k}^{*}(x)+\pi_{n}(x)\right|^{s} d x \tag{3.11}
\end{equation*}
$$

are uniformly bounded. Finally,

$$
\begin{equation*}
\lim _{k \rightarrow \infty} \int_{D_{k_{1}}}\left|p_{0}(x)-p_{k}^{*}(x)\right|^{s} d x=0 \tag{3.12}
\end{equation*}
$$

Hence, using (3.7)-(3.12), we may choose a $k$ so large that

$$
\begin{align*}
& \int_{D_{k_{1}}}\left|f\left[x, z_{0}(x), p_{0}(x)\right]-f\left[x_{k}^{*}(x), z_{k}^{*}(x), p_{k}^{*}(x)\right]\right| d x<\epsilon / 5,  \tag{3.13}\\
& \int_{D_{k_{1}}}\left|f\left[x, z_{0}(x), p_{n}(x)\right]-f\left[x_{k}^{*}(x), z_{k}^{*}(x) p_{k}^{*}(x)+\pi_{n}(x)\right]\right| d x<\epsilon / 5 \\
& (n=1,2, \cdots),
\end{align*}
$$

and then choose $n_{1}$ so large that

$$
\begin{equation*}
\int_{D_{k_{1}}}\left|f\left[x, z_{n}(x), p_{n}(x)\right]-f\left[x, z_{0}(x), p_{n}(x)\right]\right| d x<\epsilon / 5, \quad n>n_{1} . \tag{3.15}
\end{equation*}
$$

Since $x_{k}^{*}(x), z_{k}^{*}(x), p_{k}^{*}(x)$ are constant on each $R_{k i}$, it follows from Lemnia 3.4 that

$$
\begin{equation*}
\liminf _{n \rightarrow \infty} \int_{D_{k_{1}}} f\left[x_{k}^{*}(x), z_{k}^{*}(x), p_{k}^{*}(x)+\pi_{n}(x)\right] d x \tag{3.16}
\end{equation*}
$$

$$
\geqq \int_{D_{k_{1}}} f\left[x_{k}^{*}(x), z_{k}^{*}(x), p_{k}^{*}(x)\right] d x .
$$

Using (3.6) and (3.13)-(3.16), we see that

$$
\liminf _{n \rightarrow \infty} I\left(z_{n}, D\right) \geqq I\left(z_{0}, D\right)-\epsilon
$$

The result follows in this case.
We now consider the alternative (b). For each natural number $q$, we define

$$
\begin{aligned}
& \quad f_{q}(x, z, p)=\left[1-a_{q}(x, z)\right] f(x, z, p)+k \cdot a_{q}(x, z), \\
& a_{q}(x, z)= \begin{cases}0 & (0 \leqq R \leqq q) \\
3(R-q)^{2}-2(R-q)^{3} & (q \leqq R \leqq q+1), \\
1 & (R \geqq q+1), \quad R=\left(|x|^{2}+|z|^{2}\right)^{1 / 2}\end{cases}
\end{aligned}
$$

Remembering (3.3)-(3.5), we see that each $f_{q}$ satisfies hypotheses (i)-(iv) with the same $k$ and some $K_{q}$. Moreover $f_{q}$ is independent of $(x, z)$ for $R \geqq q+1$, and also

$$
f_{q}(x, z, p) \leqq f_{q+1}(x, z, p), \lim _{q \rightarrow \infty} f_{q}(x, z, p)=f(x, z, p)
$$

Thus it is sufficient to prove the lower semicontinuity for each $q$.
For a fixed $q$, we note that we may replace $\left|z_{0}(x)-z_{k}^{*}(x)\right|$ by $\phi_{k}(x)$ in (3.7) and (3.8) and $\left|z_{n}(x)-z_{0}(x)\right|$ by $\psi_{n}(x)$ in (3.9), where

$$
\begin{aligned}
& \phi_{k}(x)=\min \left(\left|z_{0}(x)-z_{k}^{*}(x)\right|, 2 q+2\right) \\
& \psi_{n}(x)=\min \left(\left|z_{n}(x)-z_{0}(x)\right|, 2 q+2\right)
\end{aligned}
$$

From the uniform boundedness of the $\phi_{k}$ and $\psi_{n}$ ( $q$ fixed), the uniform absolute continuity of the set function $D_{s}\left(z_{n}, e\right)$, and the facts that

$$
\lim _{k \rightarrow \infty} \phi_{k}(x)=0, \lim _{n \rightarrow \infty} \psi_{n}(x)=0
$$

almost everywhere, it follows that the argument can be carried through as before for each fixed $q$.

Theorem 3.2. Suppose $s>\nu$ and suppose $f$ satisfies the hypotheses of Theorem 3.1 with (i) replaced by

$$
\begin{equation*}
f(x, z, p) \geqq m|p|^{s}+k \quad(m>0) \tag{í}
\end{equation*}
$$

If $z^{*}$ is any function of class $\Re_{s}$ on the bounded domain $D$, then there is a function $z_{0}$ of class $\Re_{s}$ which coincides with $z^{*}$ on $D^{*}$ and minimizes $I(z, D)$ among all such functions.

Proof. Let $z_{n}$ be a minimizing sequence. It follows from ( $\mathrm{i}^{\prime}$ ) that $D_{s}\left(z_{n}, D\right)$ is uniformly bounded. From [2], Theorem 9.4, it follows that $\bar{D}_{s}\left(z_{n}, D\right)$ is uniformly bounded. But then a subsequence, still called $\left\{z_{n}\right\}$, converges weakly in $\Re_{s}$ to some function $z_{0}$ of class $\Re_{s}$ which coincides with $z^{*}$ on $D^{*}$ by [2], Theorem 9.2. But, from [3], Chapter II, Theorem 2.1, it follows that the equivalent functions $\bar{z}_{n}$ and $\bar{z}_{0}$ are equicontinuous on closed sets interior to $D$. Hence $z_{n}$ converges uniformly to $\bar{z}_{0}$ on each closed set interior to $D$. Hence, from the preceding theorem, $z_{0}$ is a desired solution.

More general theorems involving variable boundary values, similar to those in [3], Chapter III, §5, with $s>\nu$, can be proved.
4. Necessary conditions for quasi-convexity. In the two preceding sections, we have established the connection between quasi-convexity and lower semicontinuity. In this section, we shall establish some necessary conditions for quasi-convexity. In the next section, we establish some sufficient conditions which are also necessary when $f$ has certain interesting special forms. Unfortunately, the writer is unable to establish conditions which are both necessary and sufficient in the general case.

Lemma 4.1. Suppose $f$ is continuous, $Q$ is the cell

$$
\left|x^{\alpha}\right| \leqq 1 \quad(\alpha=1, \cdots, \nu), \delta>0,
$$

and suppose

$$
\begin{equation*}
\int_{Q} f[p+\pi(x)] d x \geqq f(p) \cdot m(Q) \tag{4.1}
\end{equation*}
$$

for every function $\zeta$ which satisfies a Lipschitz condition with coefficient $<\delta$ on $\bar{Q}$ and vanishes on $Q^{*}$. Then (4.1) also holds with $Q$ replaced by any bounded domain $D$.

Proof. Suppose $\zeta$ satisfies the conditions on the bounded domain $\bar{D}$. Let $\bar{R}$ be a hypercube of side $h$ which contains $\bar{D}$, and extend $\zeta$ to $\bar{R}$ :

$$
x_{0}^{\alpha} \leqq x^{\alpha} \leqq x_{0}^{\alpha}+h
$$

by defining $\zeta=0$ on $\bar{R}-\bar{D}$. Then $\zeta$ satisfies the conditions on $\bar{R}$, and

$$
\zeta^{*}(x)=h^{-1} \zeta\left(x_{0}+h x\right)
$$

satisfies the conditions on $\bar{Q}$, and

$$
\zeta_{x^{a}}^{* i}(x)=\zeta_{x^{\alpha}}^{i}\left(x_{0}+h x\right)
$$

Definition 4.1. The function $f$ is said to be weakly quasi-convex if with each $p$ is associated a $\delta_{p}>0$ such that (4.1) holds for all $D$ and all $\zeta$ satisfying a Lipschitz condition with coefficient $<\delta_{p}$ and vanishing on $D^{*}$.

In other words, $f$ is weakly quasi-convex if and only if each linear function furnishes a weak relative minimum among all Lipschitzian functions coinciding with it on the boundary, whereas $f$ is quasi-convex if and only if any linear function furnishes the absolute minimum among all such functions. Thus we have the following result.

Theorem 4.l. If $f$ is continuous and quasi-convex, it is weakly quasiconvex.

We shall see that if $f$ is weakly quasi-convex and continuous, then $f$ satisfies a uniform Lipschitz condition on any bounded set in $p$-space and satisfies a generalized Weierstrass condition (see Theorem 4.3) which reduces to the ordinary Weierstrass condition if $f$ is of class $C^{\prime}$ (see (4.7)) and is equivalent to the Legendre-Hadamard condition (see (4.8)) if $f$ is of class $C^{\prime \prime}$.

Lemma 4.2. Suppose $\phi$ is continuous, and suppose corresponding to any point $\lambda_{0}$ in $E_{\nu}$ there is a $\delta>0$ such that for any unit vector $\mu$ we have

$$
k \phi\left(\lambda_{0}-h \mu\right)+h \phi\left(\lambda_{0}+k \mu\right) \geqq(h+k) \phi\left(\lambda_{0}\right) \quad(0<h<\delta, 0<k<\delta) .
$$

Then $\phi$ is convex in $\lambda$.
Proof. Let $\lambda_{0}$ be any point, and $\mu$ any point with $|\mu|=1$. We shall show that

$$
\psi(t)=\phi\left(\lambda_{0}+\mu t\right)
$$

is convex in $t$. From the hypothesis, it follows that for each $t_{0}$, there is a $\delta\left(t_{0}\right)>0$ such that
(4.2) $k \psi\left(t_{0}-h\right)+h \psi\left(t_{0}+k\right) \geqq(h+k) \psi\left(t_{0}\right) \quad(0 \leqq h \leqq \delta, \quad 0 \leqq k \leqq \delta)$.

Now, suppose $t_{1}<t_{2}$. Let

$$
\chi(t)=\psi(t)-\psi\left(t_{1}\right)-\frac{t-t_{1}}{t_{2}-t_{1}}\left[\psi\left(t_{2}\right)-\psi\left(t_{1}\right)\right] .
$$

Then $\chi(t)$ satisfies (4.2) and $\chi\left(t_{1}\right)=\chi\left(t_{2}\right)=0$. Suppose $M=\max \chi(t)$ ( $t_{1} \leqq t \leqq t_{2}$ ), and suppose $M>0$. Let $t_{0}$ be the smallest value of $t$ such that $\chi(t)=M$, and let the number $\delta\left(t_{0}\right)$ be chosen as above. Clearly $t_{1}<t_{0}<t_{2}$. Choose $t_{3}$ and $t_{4}$ with

$$
\left|t_{3}-t_{0}\right|<\delta,\left|t_{4}-t_{0}\right|<\delta \quad\left(t_{1} \leqq t_{3}<t_{0}<t_{4} \leqq t_{2}\right)
$$

Then $\chi\left(t_{3}\right)<M, \quad \chi\left(t_{4}\right) \leqq M$, so that

$$
\left(t_{4}-t_{0}\right) \chi\left[t_{0}-\left(t_{0}-t_{3}\right)\right]+\left(t_{0}-t_{3}\right) \chi\left[t_{0}+\left(t_{4}-t_{0}\right)\right]<\left(t_{4}-t_{3}\right) \chi\left(t_{0}\right),
$$

which contradicts the hypothesis. Thus $\chi(t) \leqq 0$, so that

$$
\psi(t) \leqq \psi\left(t_{1}\right)+\frac{t-t_{1}}{t_{2}-t_{1}} \quad\left[\psi\left(t_{2}\right)-\psi\left(t_{1}\right)\right] .
$$

Since $t_{1}$ and $t_{2}$ were arbitrary with $t_{1}<t_{2}$, the function $\psi$ is convex in $t$. Thus $\phi$ is convex in $\lambda$.

THEOREM 4.2. If $f$ is weakly quasi-convex, then $f\left(p_{\alpha}^{i}+\lambda_{\alpha} \xi^{i}\right)$ is convex in $\lambda$ for each fixed $p$ and $\xi$.

Proof. Let $p_{\alpha}^{i}, \xi^{i}$ and $\lambda_{0 \alpha}$ be fixed and let $\mu_{1}$ be any unit vector, and suppose $h>0, k>0$. Choose $\delta\left(p_{\alpha}^{i} ; \xi^{i}, \lambda_{0 \alpha}\right)>0$ but so small that, for any bounded domain $G$,

$$
\begin{equation*}
\int_{G} f\left[p_{\alpha}^{i}+\lambda_{0 \alpha} \xi^{i}+\zeta_{x^{\alpha}}^{i}(x)\right] d x \geqq m(G) f\left(p_{\alpha}^{i}+\lambda_{0 \alpha} \xi^{i}\right) \tag{4.3}
\end{equation*}
$$

for all $\zeta$ satisfying a Lipschitz condition of constant $<\delta$ on $G$ and vanishing on $G^{*}$. Let $\left(\mu_{1}, \cdots, \mu_{\nu}\right)$ be a normal orthogonal set of unit vectors. If $\xi=0$, the result is obvious. If $\xi \neq 0$, choose $h$ and $k$ with $0<h|\xi|<\delta, 0<k|\xi|<\delta$, and let $\rho$ be any number $>|\xi| / \delta$. Let $H=(1 / \rho) k, K=(1 / \rho) h$, and let $R$ be the rectangular parallelepiped

$$
-\rho H \leqq y^{1} \leqq \rho K,\left|y^{\beta}\right| \leqq \rho \quad(\beta=2, \cdots, \nu)
$$

where

$$
y^{\beta}=x, \mu_{\beta}
$$

Let $F_{1}^{-}$be the face $y^{1}=-\rho H, F_{1}^{+}$be the face $y^{1}=\rho K, F_{\beta}^{-}$be the face $y^{\beta}=-\rho, F_{\beta}^{+}$be the face $y^{\beta}=\rho$, and let $\Pi_{\beta}^{-}$and $\Pi_{\beta}^{+}$be the pyramids with vertex at the origin and base $F_{\beta}^{-}$and $F_{\beta}^{+}$, respectively. Let $\zeta$ be defined on $\bar{R}$ to be continuous on $\bar{R}$, zero on $R^{*}$, linear on each $\Pi_{\beta}$ and $\Pi_{\beta}^{+}$, with $\zeta(0)=\xi$. Then

$$
\zeta_{x^{\alpha}}^{i}= \begin{cases}(\rho H)^{-1} \mu_{1 \alpha} \xi^{i}=k_{\mu_{1 \alpha}} \xi^{i} & , \text { on } \Pi_{1}^{-}  \tag{4.4}\\ -(\rho K)^{-1} \mu_{1 \alpha} \xi^{i}=-h_{\mu_{1 \alpha}} \xi^{i}, & \text { on } \Pi_{1}^{+} \\ \rho^{-1} \mu_{\beta \alpha} \xi^{i} & , \text { on } \Pi_{\beta}^{-} \\ -\rho^{-1} \mu_{\beta \alpha} \xi^{i} & , \text { on } \Pi_{\beta}^{+}\end{cases}
$$

Also

$$
m\left(\Pi_{1}^{-}\right)=\nu^{-1} 2^{\nu-1} \rho^{\nu} H, m\left(\Pi_{1}^{+}\right)=\nu^{-1} 2^{\nu-1} \rho^{\nu} K, m(R)=2^{\nu-1} \rho^{\nu}(H+K)
$$

$$
\begin{equation*}
m\left(\mathrm{II}_{\beta}^{-}\right)=m\left(\mathrm{II}_{\beta}^{+}\right)=\nu^{-1} 2^{\nu-2} \rho^{\nu}(H+K) \quad(\beta=2, \cdots, \nu) \tag{4.5}
\end{equation*}
$$

Then, by applying (4.3), (4.4), and (4.5), we obtain

$$
\begin{aligned}
& \frac{1}{2 \nu}\left\{\frac{2 k}{h+k} f\left[p_{\alpha}^{i}+\left(\lambda_{0 \alpha}-h \mu_{1 \alpha}\right) \xi^{i}\right]+\frac{2 h}{h+k} f\left[p_{\alpha}^{i}+\left(\lambda_{0 \alpha}+k \mu_{1 \alpha}\right) \xi^{i}\right]\right. \\
& \left.+\sum_{\beta=2}^{\nu}\left(f\left[p_{\alpha}^{i}+\left(\lambda_{0 \alpha}-\rho^{-1} \mu_{\beta \alpha}\right) \xi^{i}\right]+f\left[p_{\alpha}^{i}+\left(\lambda_{0 \alpha}+\rho^{-1} \mu_{\beta \alpha}\right)\right]\right)\right\} \\
& \geqq f\left(p_{\alpha}^{i}+\lambda_{0 \alpha} \xi^{i}\right)
\end{aligned}
$$

Letting $\rho \longrightarrow \infty$, we obtain

$$
k f\left[p_{\alpha}^{i}+\left(\lambda_{0 \alpha}-h \mu_{1 \alpha}\right) \xi^{i}\right]+h f\left[p_{\alpha}^{i}+\left(\lambda_{0 \alpha}+k \mu_{1 \alpha}\right) \xi^{i}\right] \geqq(h+k) f\left(p_{\alpha}^{i}+\lambda_{0 \alpha} \xi^{i}\right)
$$

From the preceding lemma, it follows that $f\left(p_{\alpha}^{i}+\lambda_{\alpha} \xi^{i}\right)$ is convex in $\lambda$ for each $\xi$ and $p$.

Theorem 4.3. Suppose $f$ is continuous and convex in $\lambda$ for $p$ and $\xi$.
Then $f$ satisfies a uniform Lipschitz condition on each bounded closed set, and for each fixed $p$ there exists a set of constants $A_{i}^{\alpha}$ such that

$$
\begin{equation*}
f\left(p_{\alpha}^{i}+\lambda_{\alpha} \xi^{i}\right) \geqq f(p)+A_{i}^{\alpha} \lambda_{\alpha} \xi^{i} \tag{4.6}
\end{equation*}
$$

for all $\lambda$ and $\xi$. If $f$ is of class $C^{\prime}$, (4.6) holds if and only if $A_{i}^{a}=f_{p_{a}^{i}}$, that is,

$$
\begin{equation*}
f\left(p_{\alpha}^{i}+\lambda_{\alpha} \xi^{i}\right) \geqq f(p)+f_{p_{\alpha}^{i}}(p) \lambda_{\alpha} \xi^{i} . \tag{4.7}
\end{equation*}
$$

If $f$ is of class $C^{\prime \prime},(4.7)$ holds for all $p, \lambda, \xi$ if and only if

$$
\begin{equation*}
f_{p_{\alpha}^{i} p_{\beta}^{j}}(p) \lambda_{\alpha} \lambda_{\beta} \xi^{i} \xi^{j} \geqq 0 \tag{4.8}
\end{equation*}
$$

for all $\lambda, \xi, p$.
Proof. Suppose, first, that $f$ is of class $C^{\prime}$. Let $p$ and $\xi$ be fixed. Then (4.7) follows from the convexity in $\lambda$. Moreover, since each unit vector $e_{\alpha}^{i}$ in the $p$-space is of the form $\lambda_{\alpha} \xi^{i}$, we see from the convexity in $\lambda$ that

$$
\begin{equation*}
f(p)-f\left(p-e_{\alpha}^{i}\right) \leqq f_{p_{\alpha}^{i}}(p) \leqq f\left(p+e_{\alpha}^{i}\right)-f(p) \tag{4.9}
\end{equation*}
$$

for all $p$. Thus the derivatives of $f$ are uniformly bounded by these differences in the values of $f$ on any bounded part of space. Moreover, in this case, if constants $A_{i}^{\alpha}$ satisfy (4.6), we must have

$$
A_{i}^{\alpha}=f_{p_{a}^{i}}(p)
$$

Now, if $f$ is of class $C^{\prime \prime}$, equation (4.8) with $p$ replaced by $p_{\alpha}^{i}+\lambda_{\alpha} \xi^{i}$ is equivalent to the condition that $f$ is convex in $\lambda$ for each fixed $p$ and $\xi$.

Finally, if $f$ is continuous and has this stated convexity property, it is clear that the $h$-average function also does, and $f_{h}$ is of class $C^{\prime}$. By letting $h \rightarrow 0$, we see that $f$ satisfies a uniform Lipschitz condition on any bounded closed set. Now, choose $h_{n}=n^{-1}$ and choose $p$ fixed. From (4.9) and the uniform convergence of $f_{h}$ to $f$ on any bounded part of space, we conclude that the derivatives $f_{h_{n} p_{a}^{i}}(p)$ are uniformly bounded. We may therefore choose a subsequence, still called $h_{n}$, such that

$$
\lim _{n \rightarrow \infty} f_{h_{n} p_{\alpha}^{i}}(p)=A_{i}^{\alpha} .
$$

Since (4.7) holds for all $\lambda$ and $\xi$ for each $n$, (4.6) holds in the limit.
5. Sufficient conditions for quasi-convexity. In this section we prove one general sufficient condition and then give conditions which are necessary and sufficient when $f$ has certain interesting special forms.

Lemma 5.1. Suppose $\zeta$ satisfies a uniform Lipschitz condition on the closure $\bar{D}$ of the bounded domain $D$ and suppose $\zeta=0$ on $D^{*}$. If

$$
1 \leqq \mu \leqq \nu, 1 \leqq i_{1}, \cdots, i_{\mu} \leqq N, 1 \leqq \alpha_{1}<\alpha_{2}<\cdots<\alpha_{\mu} \leqq \nu
$$

then

$$
\int_{D} \frac{\partial\left(\zeta^{i_{1}}, \cdots, \zeta^{i_{\mu}}\right)}{\partial\left(x^{a_{1}}, \ldots, x^{\alpha_{\mu}}\right)} d x=0
$$

Proof. Choose a large cell $R$ containing $\bar{D}$ in its interior, and extend $\zeta$ by defining it to be zero outside $\bar{D}$. Then the second $h$-average function $\zeta_{h h}$ is of class $C^{\prime \prime}$ on $\bar{R}$ and vanishes on and near $R^{*}$. Since any integral of the above type formed for $\zeta_{h h}$ tends to that for $\zeta$ as $h \rightarrow 0$, we need prove the theorem only for functions $\zeta$ of class $C^{\prime \prime}$ on cells $\bar{R}$.

As an example, take $i_{\beta}=\alpha_{\beta}=\beta, \beta=1, \cdots, \mu, D=R$. Then

$$
\begin{aligned}
\int_{R} \frac{\partial\left(\zeta^{1}, \cdots, \zeta^{\mu}\right)}{\partial\left(x^{1}, \ldots, x^{\mu}\right)} d x & =\int_{R} \sum_{a=1}^{\mu}(-1)^{\mu+\alpha} \zeta_{x^{\alpha}}^{\mu} Q d x \\
& =\int_{R^{*}} \zeta^{\mu} \sum_{a=1}^{\mu}(-1)^{\mu+\alpha} Q d x_{a}^{\prime} \\
& -\int_{R}(-1)^{\mu} \zeta^{\mu} \sum_{a=1}^{\mu}(-1)^{\alpha} \frac{\partial}{\partial x^{\alpha}} Q d x
\end{aligned}
$$

where

$$
Q=\frac{\partial\left(\zeta^{1}, \cdots, \zeta^{\alpha-1}, \zeta^{\alpha}, \cdots, \zeta^{\mu-1}\right)}{\partial\left(x^{1}, \cdots, x^{\alpha-1}, x^{\alpha+1}, \cdots, x^{\mu}\right)}
$$

the last equality holding by Green's theorem. But the boundary integral vanishes since $\zeta=0$ on $R^{*}$, and the integrand in the second integral vanishes on $R$ (see [3], Chapter II, Lemma 1.1, for instance).

Theorem 5.1. A sufficient condition that $f$ be quasi-convex is that for each $p$ there exist alternating forms

$$
A_{i}^{a} \pi_{a}^{i}, A_{i j}^{\alpha \beta} \pi_{\alpha}^{i} \pi_{\beta}^{j}, \cdots, A_{i_{1}}^{\alpha_{1}}, \cdots,,_{\nu} \pi_{a_{1}}^{i_{\nu}} \cdots \pi_{a_{\nu}}^{i_{\nu}}
$$

such that for all $\pi$ we have

$$
f(p+\pi) \geqq f(p)+A_{i}^{\alpha} \pi_{a}^{i}+\cdots+A_{i_{1}}^{a_{1}}, \cdots, a_{\nu} \pi_{a_{1}}^{i_{1}} \cdots \pi_{\alpha_{\nu}}^{i_{\nu}} .
$$

Proof. This is an immediate consequence of the preceding lemma.
Theorem 5.2. If the $a_{i j}^{a \beta}$ are constants and

$$
\begin{equation*}
f(p)=a_{i j}^{\alpha \beta} p_{\alpha}^{i} p_{\beta}^{j}, \tag{5.1}
\end{equation*}
$$

a necessary and sufficient condition that $f$ be quasi-convex is that

$$
\begin{equation*}
a_{i j}^{\alpha \beta} \lambda_{\alpha} \lambda_{\beta} \xi^{i} \xi^{j} \geqq 0 \tag{5.2}
\end{equation*}
$$

for all $\lambda$ and $\xi$.
Proof. If $\zeta=0$ on $D^{*}$, we see from Lemma 5.1 that

$$
\int_{D} f[p+\pi(x)] d x=f(p) m(D)+\int_{D} a_{i j}^{\alpha \beta} \pi_{\alpha}^{i}(x) \pi_{\beta}^{j}(x) d x
$$

But Van Hove [6] has shown that the condition (5.2) is necessary (this also follows from Theorem 4.3) and sufficient for the second integral to be $\geqq 0$ for all $\zeta$ of class $D^{\prime}$ on $D$ which vanish on $D^{*}$ (hence this is true also for all $\zeta$ of class $\beta_{2}$ on $D$ and vanishing on $D^{*}$ ).

Lemma 5.2. Suppose

$$
\sum_{i, j=1}^{n} a_{i j} x^{i} y^{j}=0
$$

for all $x$ and $y$ for which

$$
\sum_{i, j=1}^{n} b_{i j} x^{i} y^{j}=0
$$

Then there is a constant $K$ such that

$$
a_{i j}=K b_{i j}(i, j=1, \cdots, n) .
$$

Proof. We may introduce new variables $\xi$ and $\eta$ by

$$
x=c \xi, \quad y=d \eta,
$$

$c$ and $d$ being $n \times n$ nonsingular matrices. Let $a$ and $b$ be the matrices of the original forms and $A$ and $B$ those of the transformed forms. Then

$$
A=c^{\prime} a d, \quad B=c^{\prime} b d \quad\left(c_{i j}^{\prime}=c_{j i}\right) .
$$

We shall show that there is a scalar $K$ such that $A=K B$. We may assume that

$$
B_{i i}=1 \quad(i=1, \cdots, r) ; \quad B_{i j}=0 \text { otherwise }, \quad r \leqq n,
$$

unless $B=0$ in which case $A=0$ also and the theorem holds. By taking $\eta^{s}=1$, $\eta^{j}=0 \quad(j \neq s, s=1, \cdots, n)$ in turn we see that
$A_{i s}=0(i=1, \cdots, n, s>r) ; A_{i s}=0(i \neq s, s=1, \cdots, r, i=1, \cdots, n)$.
Then, by choosing $1 \leqq s<t \leqq r$ and setting $\eta^{s}=\eta^{t}=1, \eta^{j}=0, j \neq s, j \neq t$, we have

$$
\left(A_{i s}+A_{i t}\right) \xi^{i}=0 \quad \text { for all } \xi \text { with } \xi^{s}+\xi^{t}=0
$$

Thus there exists a constant $K(s, t)$ such that

$$
A_{s s}+A_{s t}=K(s, t), \quad A_{t s}+A_{t t}=K(s, t) .
$$

Hence

$$
A_{11}=A_{22}=\cdots=A_{T r}=K,
$$

so that $A=K B$.
Theorem 5.3. Suppose that $N=\nu+1$ and

$$
\begin{equation*}
f(p)=F\left(X_{1}, \cdots, X_{\nu+1}\right), \tag{5.3}
\end{equation*}
$$

uhere $F$ is positively homogeneous of the first degree in the $X_{i}$ and

$$
\begin{aligned}
& X_{i}=-\operatorname{det} M_{i} \quad(i=1, \cdots, \nu), \quad X_{\nu+1}=\operatorname{det} M_{\nu+1}, \\
& M_{\nu+1}=\left\|p_{\alpha}^{1}, \cdots, p_{a}^{\nu}\right\|, M_{i}=\left\|p_{\alpha}^{1}, \cdots, p_{a}^{i-1}, p_{\alpha}^{\nu+1}, p_{\alpha}^{i+1}, \cdots, p_{\alpha}^{\nu}\right\|
\end{aligned}
$$

$$
(i=1, \cdots, \nu) .
$$

Then $f$ is quasi-convex in $p$ if and only if $F$ is convex in the $X_{i}$.
Proof. If $F$ is convex in the $X_{i}$, it follows from Theorem 5.1 that $f$ is quasiconvex in $p$.

Hence suppose $f$ is given by (5.3) and is quasi-convex in $p$. If

$$
\Delta X_{k}=X_{k}\left(p_{\alpha}^{i}+\lambda_{\alpha} \xi^{i}\right)-X_{k}\left(p_{\alpha}^{i}\right)
$$

then

$$
\begin{equation*}
\Delta X_{k}=X_{k p_{\alpha}^{i}} \lambda_{\alpha} \xi^{i} \tag{5.4}
\end{equation*}
$$

Also, since

$$
p_{\beta}^{k} X_{k}=0 \quad(\beta=1, \cdots, \nu)
$$

we have

$$
\begin{equation*}
\mathrm{p}_{\beta}^{k} X_{k p_{\alpha}^{i}}=-\delta_{\beta}^{a} X_{i} \tag{5.5}
\end{equation*}
$$

Now, choose a set of $X_{i}$ not all zero and choose any $p$ such that

$$
X_{i}(p)=X_{i}
$$

Since $f$ is quasi-convex and hence weakly so, there are constants $A_{\alpha}^{i}$ such that

$$
f\left(p_{a}^{i}+\lambda_{\alpha} \xi^{i}\right) \geqq f(p)+A_{i}^{a} \lambda_{\alpha} \xi^{i}
$$

Since $f$ depends only on the $X_{i}$, we must have

$$
\begin{equation*}
A_{i}^{\alpha} \lambda_{\alpha} \xi^{i} \leqq 0 \text { for all } \lambda, \xi \text { with } X_{k p_{\alpha}^{i}} \lambda_{\alpha} \xi^{i}=0(k=1, \cdots, \nu+1) \tag{5.6}
\end{equation*}
$$

Obviously, then, the equality must hold in (5.6). Using (5.4) and (5.5), we see that

$$
\begin{equation*}
p_{\beta}^{k} \Delta X_{k}=-\lambda_{\beta}\left(X_{i} \xi^{i}\right) \tag{5.7}
\end{equation*}
$$

$$
(\beta=1, \cdots, \nu)
$$

Hence, we must have

$$
\begin{equation*}
A_{i}^{a} \lambda_{\alpha} \xi^{i}=0 \tag{5.8}
\end{equation*}
$$

for all $\lambda, \xi$ for which

$$
\begin{equation*}
X_{i} \xi^{i}=0 \text { and } D_{i}^{\alpha} \lambda_{\alpha} \xi^{i}=0, D_{i}^{\alpha}=X_{k} X_{k p_{\alpha}^{i}} \tag{5.9}
\end{equation*}
$$

Now, since not all the $X_{i}$ are zero, assume $X_{k} \neq 0$. Then

$$
\begin{equation*}
\sum_{i \neq k}\left(A_{i}^{a} X_{k}-A_{k}^{\alpha} X_{i}\right) \lambda_{a} \xi^{i}=0 \tag{5.10}
\end{equation*}
$$

for all $\lambda, \xi$ for which

$$
\begin{equation*}
\sum_{i \neq k}\left(D_{i}^{\alpha} X_{k}-D_{k}^{\alpha} X_{i}\right) \lambda_{\alpha} \xi^{i}=0 \tag{5.11}
\end{equation*}
$$

From the preceding lemma, it follows that there is a constant $K$ such that

$$
\begin{equation*}
A_{i}^{\alpha} X_{k}-A_{k}^{\alpha} X_{i}=K\left(D_{i}^{\alpha} X_{k}-D_{k}^{\alpha} X_{i}\right) \tag{5.12}
\end{equation*}
$$

Hence

$$
\begin{equation*}
A_{i}^{\alpha}=K D_{i}^{\alpha}+L^{\alpha} X_{i}, \quad L^{\alpha}=X_{k}^{-1}\left(A_{k}^{\alpha}-K D_{k}^{\alpha}\right) \tag{5.13}
\end{equation*}
$$

From (5.7) and (5.13) it follows that

$$
\begin{equation*}
A_{i}^{\alpha} \lambda_{\alpha} \xi^{i}=K D_{i}^{\alpha} \lambda_{\alpha} \xi^{i}+L^{\alpha} \lambda_{\alpha} X_{i} \xi^{i}=C^{k} \Delta X_{k}, C^{k}=\left(K X_{k}-L^{\alpha} p_{\alpha}^{k}\right) \tag{5.14}
\end{equation*}
$$

Finally, if we are given any values of the $\Delta X_{k}$, the quantities

$$
h_{i}=p_{i}^{k} \Delta X_{k}(i=1, \cdots, \nu) \text { and } h_{\nu+1}=X_{k} \Delta X_{k}
$$

are determined and the $\Delta X_{k}$ are also uniquely determined by the $h_{i}$. Using (5.7), we may determine the $\lambda_{\alpha}$ in terms of the $h_{i}(i=1, \cdots, \nu)$, and substitute them into

$$
h_{\nu+1}=X_{k} \Delta X_{k}=D_{i}^{\alpha} \lambda_{\alpha} \xi^{i}
$$

and we merely have to choose the $\xi^{i}$ to satisfy the equation

$$
\left(D_{i}^{\alpha} h_{\alpha}+h_{\nu+1} X_{i}\right) \xi^{i}=0 \text { with } X_{i} \quad \xi^{i} \neq 0
$$

this is always possible unless all the $D_{i}^{\alpha} h_{\alpha}=0$. Thus, unless these linear relations in the $\Delta X_{i}$ hold, we have

$$
\begin{equation*}
F(X+\Delta X)=f\left(p_{a}^{i}+\lambda_{\alpha} \xi^{i}\right) \geqq f(p)+A_{i}^{\alpha} \lambda_{\alpha} \xi^{i}=F(X)+C^{k} \Delta X_{k} \tag{5.15}
\end{equation*}
$$

The result follows in this case by continuity.

Finally, since $F$ is homogeneous of the first degree, we see by taking

$$
\Delta X=h X, \quad h>-1
$$

that

$$
F[(1+h) X]=(1+h) F(X) \geqq F(X)+h C^{k} X_{k}
$$

or

$$
h\left[F(X)-C^{k} X_{k}\right] \geqq 0, h>-1 .
$$

Hence $F(X)=C^{k} X_{k}$. Then by setting $X=h X_{0}, X_{0} \neq 0$, choosing the $C^{k}$ for this $X_{0}$, and then letting $h \rightarrow 0$, we see that (5.15) holds for some $C^{k}$ even if $X=0$.

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# SOME INEQUALITIES IN CERTAIN NONORIENTABLE RIEMANNIAN MANIFOLDS 

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1. Introduction. As is well known, the projective plane and the Moebius strip are nonorientable manifolds of dimension two. After introducing a Riemannian metric on each of them, we obtain two 2-dimensional nonorientable Riemannian manifolds. For convenience of reference, let us denote them by $M_{p^{2}}^{2}$ and $M_{m}^{2}$, respectively. Each of these manifolds has an area $A$. Moreover, there exists a family of closed curves, which are not homotopic to zero, on each manifold; and hence the set of the lengths of all these closed curves in consideration has a positive greatest lower bound, $a$. The purpose of this paper is to investigate the relationship between these two geometrical constants, $A$ and $a$. It is found that, in each case, there exists an inequality [1] connecting them, of the form

$$
\begin{equation*}
A \geq k a^{2} \tag{1}
\end{equation*}
$$

$k$ being a constant depending only on the conformal character of the Riemannian manifold. To establish such inequalities and to determine the corresponding best possible constants are the two central problems in this investigation.

For the time being, the projective plane is used in the following realization: it is given as the unit sphere with identification of diametrically opposite points. We assume further that the metric on $M_{p^{2}}^{2}$ is given by

$$
d s^{2}=g(p) d \rho^{2},
$$

$d \rho^{2}$ being the line element of the unit sphere taken from the embedding Euclidean space; $g(p) \in C_{\omega}, g(p)>0$ for any point $p$ on the manifold. As for the Moebius strip, we assume that it is given by the strip

$$
-\beta<y<\beta,
$$

with identification given by the fundamental group

$$
\begin{aligned}
& x^{\prime}=x+n(2 \alpha), \\
& y^{\prime}=(-1)^{n} y
\end{aligned}
$$

$$
(n=0, \pm 1, \pm 2, \cdots)
$$

We assume further that the metric on $M_{m}^{2}$ is given by

$$
d s^{2}=g(x, y)\left(d x^{2}+d y^{2}\right),
$$

where $g(x, y) \in C_{\omega}$ and $g(x, y)>0$. We shall see later that these assumptions are admissible in our cases.

The main idea of the method for solving these problems is to reduce the general metric, $g(p) d \rho^{2}$, to a simple and special one, $\bar{g} d \rho^{2}$, for which the $e q$ uality in ( 1 ) holds, by an averaging process over a certain continuous group space; this enables us to handle our problems more easily. Let $A_{g}, a_{g}, A_{\bar{g}}, a_{\bar{g}}$ be the geometrical constants defined in terms of the original metric and the simplified metric respectively. Fortunately, this averaging process provides us a means of comparison between $A_{g}$ and $A_{\bar{g}}$ and between $a_{g}$ and $a_{g}$; namely, we have

$$
\begin{equation*}
A_{g} \geq A_{\bar{g}}, \tag{2}
\end{equation*}
$$

$$
a_{g} \leq a_{\bar{g}}
$$

A comparison of the equality yielded by the special metric mentioned above with the foregoing inequalities (2) gives us the desired result.

Take, for example, the manifold $M_{p 2}^{2}$. Each rotation of a 2 -sphere about its center in the ordinary space is actually a conformal mapping of $M_{p^{2}}^{2}$ onto itself. All these rotations form a compact Lie group $G$. Averaging $[g(p)] 1 / 2$ over $G$ by the Hurwitz integration,

$$
\int_{G}\left[(g(p))^{1 / 2}\right]^{\sigma} \quad \delta \sigma=h^{1 / 2}
$$

where $\sigma \in G$, and where $\delta \sigma$ is the invariant volume element, we can easily show that $h$ is a constant and that the simplified metric is an elliptic one; this produces the equality

$$
\begin{equation*}
A_{h}=\frac{2}{\pi} a_{h}^{2} \tag{3}
\end{equation*}
$$

A combination of (3) with the following inequalities corresponding to (2),

$$
\begin{aligned}
& A_{g} \geq A_{h} \\
& a_{g} \leq a_{h}
\end{aligned}
$$

shows that we have, in general,

$$
A \geq \frac{2}{\pi} a^{2}
$$

The same method can be extended, with some restrictions, to the case of $M_{p^{n}}^{n}$, that is, the Riemannian manifold whose underlying topological space is an $n$-dimensional projective space.

In the case of $M_{m}^{2}$, let the rectangle

$$
R:\left\{\begin{array}{l}
-\alpha \leq x<\alpha \\
-\beta<y<\beta
\end{array}\right.
$$

be its fundamental region, as will be explained in $\oint 3$. There exists a one-parameter family of conformal mappings of $M_{m}^{2}$ onto itself,

$$
\begin{aligned}
& x^{\prime}=x+c \\
& y^{\prime}=y
\end{aligned}
$$

$c$ being real $\bmod (4 \alpha)$. Averaging $[g(x+c, y)]^{1 / 2}$ over the interval $[0,4 \alpha]$ by the formula

$$
\frac{1}{4 \alpha} \int_{0}^{4 a}[g(x+c, y)]^{1 / 2} \quad d c=[\bar{g}(y)]^{1 / 2}
$$

we can see that $[\bar{g}(y)]^{1 / 2}$ is free of $x$ and is an even function on account of the fact that the metric is invariant under the fundamental group $\Gamma$; that is,

$$
g\left[x+n(2 \alpha),(-1)^{n} y\right]=g(x, y)
$$

A further consideration of the same problem with the metric $g(y)\left(d x^{2}+d y^{2}\right)$, where $g(y)$ is positive and even, leads to a distinguished $g_{0}(y)$ such that $g_{0}(y)\left(d x^{2}+d y^{2}\right)$ plays the same role as the elliptic metric in the case of $M_{p}^{2}$ or $M_{p^{n}}^{n}$; that is, $g_{0}(y)\left(d x^{2}+d y^{2}\right)$ leads to the equality in (1).
2. Riemannian manifold $M_{p^{2}}^{2}$, whose underlying topological space is a projective plane $P^{2}$. To begin with, let us prove the following general lemma, which will often be used.

Lemma 1. Let $M_{i}^{2}(i=1,2, \cdots, n)$ be a set of $n$ 2-dimensional Riemannian manifolds smooth of order 1 such that each $M_{i}^{2}$, with the same underlying topological space $T^{2}$, has a metric of the form

$$
d s_{i}^{2}=g_{i}(p) d s^{2}
$$

where $g_{i}(p)>0, g_{i}(p) \in C_{0}$ for $p \in M_{i}^{2}$; and $d s^{2}$ is a Riemannian metric which can be defined locally by

$$
d s^{2}=\sum_{j, k=1}^{2} g_{j k}\left(u_{1}, u_{2}\right) d u^{j} d u^{k}, \quad g_{j k}\left(u_{1}, u_{2}\right) \in C_{0}
$$

Let $\bar{g}_{n}(p)$ be defined by the formula

$$
\bar{g}_{n}(p)=\left\{\frac{\left[g_{1}(p)\right]^{1 / 2}+\cdots+\left[g_{n}(p)\right]^{1 / 2}}{n}\right\}^{2}
$$

If the sets of lengths $S_{i}=\int_{C}\left[g_{i}(p)\right]^{1 / 2} d s(i=1,2, \cdots, n)$ of a family $F$ of curves $C$ on $T^{2}$ have the same nonnegative greatest lower bound,

$$
a_{g_{1}}=a_{g_{2}}=\cdots=a_{g_{n}},
$$

and the areas $A_{g_{i}}$ of $M_{i}^{2}$ have the same value,

$$
A_{g_{1}}=A_{g_{2}}=\cdots=A_{g_{n}}
$$

then we have

$$
\begin{equation*}
A_{g_{n}}^{-} \leq A_{g_{1}}=\cdots=A_{g_{n}} \tag{i}
\end{equation*}
$$

and

$$
\begin{equation*}
a_{g_{n}} \geq a_{g_{1}}=\cdots=a_{g_{n}} \tag{ii}
\end{equation*}
$$

Proof. By the definition of area and that of $\bar{g}_{n}$, we have

$$
A \bar{g}_{n}=\iint \bar{g}_{n}(p) d \omega=\iint \frac{\left[\left(g_{1}\right)^{1 / 2}+\cdots+\left(g_{n}\right)^{1 / 2}\right]^{2}}{n^{2}} d \omega
$$

where $d \omega$ is the area element, which can be expressed locally by the formula

$$
d \omega=\left|\begin{array}{ll}
g_{11} & g_{12} \\
g_{12} & g_{22}
\end{array}\right|^{1 / 2} d u^{1} d u^{2}
$$

Making use of the inequality

$$
\left(\alpha_{1}+\cdots+\alpha_{n}\right)^{2} \leq n\left(\alpha_{1}^{2}+\cdots+\alpha_{n}^{2}\right)
$$

we have

$$
\begin{aligned}
A \bar{g}_{n} & \leq \iint n \frac{\left(g_{1}+\cdots+g_{n}\right)}{n^{2}} d \omega \\
& \leq \frac{\iint g_{1} d \omega+\cdots+\iint g_{n} d \omega}{n} \\
& \leq \frac{A_{g_{1}}+\cdots+A_{g_{n}}}{n} .
\end{aligned}
$$

By hypothesis, it follows that

$$
A \bar{g}_{n} \leq A_{g_{1}}=\cdots=A_{g_{n}},
$$

which is (i).
The proof of (ii) follows from the definitions of the concepts concerned,

$$
\begin{aligned}
\int_{C} \bar{g}_{n}^{1 / 2} d s & =\int_{C} \frac{g_{1}^{1 / 2}+\cdots+g_{n}^{1 / 2}}{n} d s=\frac{\int_{C} g_{1}^{1 / 2} d s+\cdots+\int_{C} g_{n}^{1 / 2} d s}{n} \\
& \geq \frac{a_{g_{1}}+\cdots+a_{g_{n}}}{n}=a_{g_{1}}=\cdots=a_{g_{n}},
\end{aligned}
$$

the line integrals being extended along any curve $C$ of the family $F$. Hence

$$
\begin{aligned}
& a_{\bar{g}_{n}}=\text { g.l.b. }\left(\int_{C} \bar{g}_{n}^{1 / 2} d s\right) \geq a_{g_{1}}=\cdots=a_{g_{n}} . \\
& \quad c \in F
\end{aligned}
$$

We shall now prove the following theorem, which characterizes the relationship between the two geometrical constants $A$ and $a$ in $M_{p{ }^{2}}^{2}$.

Theorem 1. Let $M_{p^{2}}^{2}$ be the Riemannian manifold whose underlying topological space is a projective plane and whose metric is locally defined by

$$
d s^{2}=\sum_{i, k=1}^{2} g_{i k}\left(u_{1}, u_{2}\right) d u^{i} d u^{k}, \quad g_{i k}\left(u_{1}, u_{2}\right) \in C_{0}
$$

let $A$ be its area, and a the greatest lower bound of the lengths of all the closed curves not homotopic to zero on $M_{p^{2}}^{2}$; then

$$
A \geq \frac{2}{\pi} a^{2}
$$

Moreover, $2 / \pi$ is the best constant.
Proof. On account of the Weierstrass approximation theorem, it suffices to assume that the fundamental tensor $g_{i k}\left(u_{1}, u_{2}\right)$ is analytic. Then we can introduce, in the small, an isothermic coordinate system on $M_{p^{2}}^{2}$ so that the metric takes the form

$$
g^{*}\left(v_{1}, v_{2}\right)\left(d v_{1}^{2}+d v_{2}^{2}\right)
$$

where $g^{*}\left(v_{1}, v_{2}\right) \in C, g^{*}\left(v_{1}, v_{2}\right)>0$. We define the metric on the universal covering surface $S^{2}$ of the projective plane by a projection process. The Riemannian manifold $M^{2}{ }_{S 2}$ thus obtained is actually a Riemannian surface. According to the uniformization theorem, we can map $M_{S^{2}}^{2}$ onto the unit 2-sphere manifold $M_{U^{2}}^{2}$, and can arrange it in such a way that two diametrically opposite points of $U^{2}$ correspond to the same point of $M_{p 2}^{2}$. The metric has then the form

$$
d s^{2}=g(p) d \rho^{2} . \quad g(p)>0, g(p) \in C_{\omega} \text { for } p \in M_{U^{2}}^{2}
$$

where $d \rho^{2}$ is, the line element of the unit sphere $U^{2}$ taken from the embedding 3-dimensional Euclidean space.

We remark that the area $A$ of $M_{p^{2}}^{2}$ is one half that of $M_{U^{2}}^{2}$.
Let us consider all the rotations $\sigma$ of the unit sphere $U^{2}$ about its center. All these rotations form a compact Lie group G. Applying the process of averaging over a compact Lie group, in this case the Hurwitz integration [2; 3, p.188], we have

$$
\int_{G}\left[(g(p))^{1 / 2}\right]^{\sigma} \delta \sigma=[h(p)]^{1 / 2} .
$$

We shall show that $[h(p)]^{1 / 2}$ is invariant with respect to all the left translations

$$
\tau: \sigma \rightarrow \sigma^{\prime}=\tau \sigma, \quad \tau, \sigma, \sigma^{\prime} \in G,
$$

and hence is a constant. In fact, let $\tau$ be any element of $G$; then, by definition,

$$
\left[(h(p))^{1 / 2}\right]^{\tau}=\int_{G}\left[(g(p))^{1 / 2}\right]^{\tau \sigma} \delta \sigma=\int_{G}\left[(g(p))^{1 / 2}\right]^{\tau \sigma} \delta \tau \sigma,
$$

since $\delta \sigma$ is invariant under all left translations. Therefore,

$$
[h(\tau p)]^{1 / 2}=\int_{G}\left[(g(p))^{1 / 2}\right]^{\lambda} d \lambda=[h(p)]^{1 / 2}, \quad \lambda=\tau_{\sigma} \in G .
$$

As the group $G$ is transitive, $h^{1 / 2}$ is a constant.
Using $h d \rho^{2}$ instead of $g d \rho^{2}$ as the metric on the unit sphere $U^{2}$, we obtain a manifold with the spherical geometry. Preserving the metric $h d \rho^{2}$, and identifying the diametrical points on $U^{2}$, we get a manifold ${ }_{h} M^{2}{ }_{p}{ }^{2}$, with the elliptic geometry. The two geometrical constants $A_{h}$ and $a_{h}$ can actually be evaluated:

$$
\begin{aligned}
A_{h} & =2 \pi h, \\
a_{h} & =\pi h^{1 / 2} .
\end{aligned}
$$

Hence

$$
\begin{equation*}
A_{h}=\frac{2}{\pi} a_{h}{ }^{2} . \tag{4}
\end{equation*}
$$

It is clear that if $g(p)$ is subjected to a transformation $\sigma$ of $G$, the resulting metric $g^{\sigma}(p) d \rho^{2}$ is such that

$$
\begin{equation*}
a_{g \sigma}=a_{g} \tag{5}
\end{equation*}
$$

and

$$
\begin{equation*}
A_{g} \sigma=A_{g} . \tag{6}
\end{equation*}
$$

By approximating integrals by suitable sums and using Lemma 1, we easily obtain

$$
\begin{equation*}
a_{h} \geq a_{g} \tag{7}
\end{equation*}
$$

and

$$
\begin{equation*}
A_{h} \leq A_{g} \tag{8}
\end{equation*}
$$

Combining (4), (7), and (8), we have

$$
A_{g} \geq A_{h}=\frac{2}{\pi} a_{h}^{2} \geq \frac{2}{\pi} a_{g}^{2}
$$

Dropping the unnecessary indices, we obtain the inequality

$$
A \geq \frac{2}{\pi} a^{2}
$$

That $2 / \pi$ is the best constant is evident, since we already have shown that the equality sign actually is attained when the metric is elliptic.

A slight generalization of Theorem 1 , referring to certain special Riemannian metrics on the $n$-dimensional projective space $P_{n}$, can be proved in a similar fashion, using Hölder's inequality

$$
\left(a_{1} b_{1}+\cdots+a_{m} b_{m}\right) \leq\left(a_{1}^{p}+\cdots+a_{m}^{p}\right)^{1 / p}\left(b_{1}^{q}+\cdots+b_{m}^{q}\right)^{1 / q}
$$

where $a_{i}, b_{i} \geq 0$ and $p, q>1$ such that $1 / p+1 / q=1$. The generalized theorem reads as follows:

Theorem 2. Let $M_{p^{n}}^{n}$ be the Riemannian manifold whose underlying topological space is an n-dimensional projective space $P^{n}$, which we suppose represented by the unit n-sphere $U^{n}$ of the $(n+1)$-dimensional Euclidean space with identification of dia:netrically opposite points $p$ and $p_{d}$, and whose metric can be represented in the form

$$
d s^{2}=g(p) d \rho^{2}
$$

where $g(p)>0, g(p) \in C_{0}, g(p)=g\left(p_{d}\right)$ for $p \in M_{p^{n}}^{n}$, and $d \rho^{2}$ is the lineelement of the n-sphere, taken from the embedding Euclidean space; let $V$ be its volume, and a the greatest lower bound of the lengths of all the closed curves which are not homotopic to zero on $M_{p n}^{n}$; then

$$
V \geq \frac{\pi^{\frac{1-n}{2}}}{\Gamma\left(\frac{n+1}{2}\right)} a^{n}=k_{n} a^{n}
$$

Further, the above $k_{n}$ is the best constant.
The proof of Theorem 2 is left to the reader.
3. Riemannian manifold $M_{m}^{2}$, whose underlying topological space is a Moebius strip. Let $M_{m}^{2}$ be the Riemannian manifold whose underlying topological space is a Moebius strip and whose metric is locally defined by

$$
d s^{2}=\sum_{i ; k=1}^{2} g_{i k}\left(u_{1}, u_{2}\right) d u^{i} d u^{k}, \quad g_{i k} \in C_{0}
$$

By Weierstrass' approximation theorem, it suffices to assume that $g_{i k} \in C_{\omega}$. After introduction of the isothermic coordinate system, the metric takes the form

$$
d s^{2}=g^{*}(u, v)\left(d u^{2}+d v^{2}\right)
$$

where $g^{*}(u, v) \in C_{\omega}$ and $g^{*}(u, v)>0$. We define the metric $d s^{-2}$ on the universal covering surface of the Moebius strip by a projection process: $d s^{2}=d \bar{s}^{2}$; that is, the metric is invariant under the fundamental group of the Moebius strip. The covering manifold of $M_{m}^{2}$ thus obtained is actually a simply connected Riemann surface. According to the uniformization theorem, we can map it conformally onto a strip

$$
S:\left\{\begin{array}{l}
-\beta<y<\beta \\
-\infty<x<\infty
\end{array}\right.
$$

of the $(x, y)$-plane. The fundamental group $\Gamma$ appears then in the form:

$$
\sigma:\left\{\begin{array}{l}
x^{\prime}=x+n(2 \alpha), \\
y^{\prime}=(-1)^{n} y
\end{array} \quad(n=0, \pm 1, \pm 2, \cdots)\right.
$$

The given manifold $M_{m}^{2}$ is mapped isogonally onto the fundamental region

$$
R:\left\{\begin{array}{l}
-\alpha \leq x<\alpha \\
-\beta<y<\beta
\end{array}\right.
$$

with a metric of the form

$$
d s^{2}=g(x, y)\left(d x^{2}+d y^{2}\right)
$$

where $g(x, y)>0$ and $g(x, y) \in C_{\omega}$. Moreover,

$$
g\left(x+n(2 \alpha),(-1)^{n} y\right)=g(x, y)
$$

We are now in a position to prove the following theorem, which connects the two geometrical constants $A$ and $a$ in $M_{m}^{2}$.

Theorem 3. Let $M_{m}^{2}$ be the Riemannian manifold whose underlying topological space is a Moebius strip and whose metric is locally defined by

$$
d s^{2}=\sum_{i, k=1}^{2} g_{i k}\left(u_{1}, u_{2}\right) d u^{i} d u^{k}, \quad g_{i k} \in C_{0}
$$

let $A$ be its area, and a the greatest lower bound of the lengths of all the closed curves which are not homotopic to zero on $M_{m}^{2}$; then we have

$$
A \geq \frac{2}{\pi} \cdot \frac{e^{\beta \pi / a}-1}{e^{\beta \pi / \alpha+1}} \cdot a^{2}=k_{\alpha \beta} a^{2}
$$

where $2 \alpha$ and $2 \beta$ are the Euclidean lengths of the sides of the fundamental region $R$ of the Moebius strip. Moreover, the above constant $k_{a \beta}$ is best for a given ratio $\beta / \alpha$.

Proof. Let us consider the continuous group

$$
H:\left\{\begin{array}{l}
x^{\prime}=x+c \\
y^{\prime}=y
\end{array}\right.
$$

$c$ being real $(\bmod 4 \alpha) ; H$ consists of conformal transformations of the Moebius strip onto itself. It is evident that every two points which are equivalent under $\Gamma$ remain equivalent under $\Gamma$ after being operated on by elements of $H$. Defining the mean value, $[\bar{g}(y)]^{1 / 2}$, of $[g(x+c, y)]^{1 / 2}$ by the formula

$$
\frac{1}{4 \alpha} \int_{0}^{4 \alpha}[g(x+c, y)]^{1 / 2} d c=[\bar{g}(y)]^{1 / 2}
$$

we can prove, by a method similar to that in the former cases, that

$$
\begin{align*}
& \bar{A} \leq A,  \tag{9}\\
& \bar{a} \geq a,
\end{align*}
$$

where $\bar{A}$ and $\bar{a}$ have the same meaning as $A$ and $a$ except that we use the metric $\bar{g}(y)\left(d x^{2}+d y^{2}\right)$ instead of the general one. Moreover, from the invariance of $g(x, y)$ under the group $\Gamma$, it follows immediately that

$$
\bar{g}(y)=\bar{g}(-y) .
$$

We now shall consider the same problem with the simple metric

$$
d s^{2}=g(y)\left(d x^{2}+d y^{2}\right)
$$

where $g(-y)=g(y)>0$. Noting (9), we see that if the best inequality is found for such a $g(y)$, it is also found for all $g(x, y)$, and hence our problem is solved.

We are now going to determine a special positive, even, and for nonnegative $y$ monotonically decreasing function $g(y)$ such that a family $F^{*}$ of closed geodesic lines through the origin and not homotopic to zero on $M_{m}^{2}$ can be defined in terms of it.

Let us first establish a differential equation for such $g(y)$. Putting

$$
d s=\left[g(y)\left(1+x^{\prime 2}\right)\right]^{1 / 2} d y \quad\left(x^{\prime}=\frac{d x}{d y}\right)
$$

we know that the equation for the extremals is

$$
\frac{d}{d y}\left\{x^{\prime}\left[\frac{g(y)}{1+x^{\prime 2}}\right]^{1 / 2}\right\}=0
$$

Solving this equation, we have

$$
x=\int_{0}^{y} \frac{c d \eta}{\left[g(\eta)-c^{2}\right]^{1 / 2}}+k
$$

Since the geodesics under consideration have to go through the origin, the constant $k$ has to be zero, and hence the equation becomes

$$
x=\int_{0}^{y} \frac{c d \eta}{\left[g(\eta)-c^{2}\right]^{1 / 2}}
$$

The condition that the geodesics of the family be closed and not homotopic to zero requires that

$$
\left.\frac{d y}{d x}\right|_{(\alpha, \tau)}=0
$$

for $-\beta<\tau<\beta$; that is,

$$
[g(\tau)]^{1 / 2}=c
$$

for $-\beta<\tau<\beta$. Hence we have

$$
\begin{equation*}
\alpha=\int_{0}^{\tau}\left[\frac{g(\tau)}{g(\eta)-g(\tau)}\right]^{1 / 2} d \eta, \quad 0 \leq \tau<\beta \tag{10}
\end{equation*}
$$

For simplicity, let us normalize $g(y)$ so that $g(0)=1$. Since $g(y)$ is supposed to be monotonically decreasing for nonnegative $y$, we can put

$$
\begin{aligned}
& 1-g(\eta)=t \\
& 1-g(\tau)=\omega
\end{aligned}
$$

and have

$$
\begin{aligned}
\omega-t & =g(\eta)-g(\tau) \\
d \eta & =-\frac{d t}{g^{\prime}(\eta)}
\end{aligned}
$$

Then equation (10) takes the form

$$
\begin{equation*}
\frac{-\alpha}{[1-\omega]^{1 / 2}}=\int_{0}^{\omega} \frac{1}{g^{\prime}(\eta)} \cdot \frac{d t}{[\omega-t]^{1 / 2}} \tag{11}
\end{equation*}
$$

This is an Abel integral equation. According to the formula (cf. [1, p.484]) for the solution of such an equation,

$$
f(x)=\int_{0}^{x} \frac{y(t) d t}{(x-t)^{1 / 2}},
$$

we have

$$
\begin{equation*}
y(t)=\frac{1}{\pi}\left[\frac{f(0)}{t^{1 / 2}}+\int_{0}^{t} \frac{f^{\prime}(z) d z}{(t-z)^{1 / 2}}\right], \tag{12}
\end{equation*}
$$

and, in our case,

$$
\frac{1}{g^{\prime}(\eta)}=-\frac{\alpha}{\pi}\left[\frac{1}{t^{1 / 2}}+\frac{1}{2} \int_{0}^{t} \frac{d z}{(1-z)^{3 / 2}(t-z)^{1 / 2}}\right]
$$

$=-\frac{\alpha}{\pi}\left[\frac{1}{t^{1 / 2}}+\frac{t^{1 / 2}}{1-t}\right]=-\frac{\alpha}{\pi} \cdot \frac{1}{(1-t) t^{1 / 2}}=-\frac{\alpha}{\pi} \cdot \frac{1}{g(\eta)[1-g(\eta)]^{1 / 2}}$
Thus we have established the differential equation for $g(y)$,

$$
\begin{equation*}
\frac{d g}{d y}=-\frac{\pi}{\alpha} g(1-g)^{1 / 2} \tag{13}
\end{equation*}
$$

The general solution of (13) is found to be

$$
\log \frac{1+(1-g)^{1 / 2}}{1-(1-g)^{1 / 2}}=\frac{\pi}{\alpha} y+k
$$

When $y=0$, we have $[1-g(0)]^{1 / 2}=0$ and hence $k=0$. Therefore,

$$
\log \frac{1+(1-g)^{1 / 2}}{1-(1-g)^{1 / 2}}=\frac{\pi}{\alpha} y
$$

An explicit expression for $g(y)$ is as follows:

$$
\begin{equation*}
g(y)=\frac{4 e^{\pi y / \alpha}}{\left(1+e^{\pi y / \alpha}\right)^{2}} \tag{14}
\end{equation*}
$$

for $-\beta<y<\beta$.
From the explicit expression (14) for $g(y)$, every property of $g(y)$ we assumed at the beginning is verified. It is a positive, even, and monotonic decreasing function for $y \geq 0$. Moreover, $g(y) \rightarrow 0$ as $y \rightarrow \infty$ and $g^{\prime}(0)=0$. Such a $g(y)$, with those properties just mentioned and defining the family $F^{*}$ of the closed geodesics through ( 0,0 ) and not homotopic to zero, is distinguished. Let us denote it by $g_{0}(y)$; that is, $g_{0}(y)$ is defined by either (13) or (14).

We are now in a position to establish the inequality in question for a positive, even, analytic function $g(y)$. By the definition of $a$, we have

$$
2 \int_{0}^{\tau}\left[g(y) \cdot\left(1+x^{\prime 2}\right)\right]^{1 / 2} d y=\varphi(\tau) \geq a
$$

for a closed curve, not homotopic to zero, on $M_{m}^{2}$. By taking this curve as one of $F^{*}$, we obtain

$$
\left(1+x^{\prime} 2\right)^{1 / 2}=\left[\frac{g_{0}(y)}{g_{0}(y)-g_{0}(\tau)}\right]^{1 / 2}
$$

Hence,

$$
\begin{equation*}
\int_{0}^{\tau}\left[\frac{g(y) g_{0}(y)}{g_{0}(y)-g_{0}(\tau)}\right]^{1 / 2} d y=\frac{\varphi(\tau)}{2} \geq \frac{a}{2} \tag{15}
\end{equation*}
$$

This equation can be put into a more suitable form by setting

$$
\begin{aligned}
& 1-g_{0}(y)=t \\
& 1-g_{0}(\tau)=\omega
\end{aligned}
$$

Then we have

$$
\begin{aligned}
\omega-t & =g_{0}(y)-g_{0}(\tau), \\
d y & =-\frac{d t}{g_{0}^{\prime}(y)}, \\
\tau & =g_{0}^{-1}(1-\omega)
\end{aligned}
$$

Equation (15) then takes the form

$$
\begin{equation*}
\int_{0}^{\omega}\left[\frac{g(y) g_{0}(y)}{\omega-t}\right]^{1 / 2} \frac{d t}{g_{0}^{\prime} \cdot(y)}=\frac{\varphi}{2}\left[g_{0}^{-1}(1-\omega)\right] \tag{16}
\end{equation*}
$$

In formula (12), we have

$$
\begin{aligned}
f(z) & =-\frac{\varphi}{2}\left[g_{0}^{-1}(1-z)\right] \\
f(0) & =-\frac{\varphi(0)}{2} \\
f^{\prime}(z) & =\frac{1}{2} \cdot \frac{\varphi^{\prime}\left(g_{0}^{-1}(1-z)\right)}{g_{0}^{\prime}\left(g_{0}^{-1}(1-z)\right)}
\end{aligned}
$$

Solving (16), we obtain

$$
\frac{\left[g(y) g_{0}(y)\right]^{1 / 2}}{g_{0}^{\prime} \cdot(y)}=\frac{1}{\pi}\left[-\frac{\varphi(0)}{2 t^{1 / 2}}+\frac{1}{2} \int_{0}^{t} \frac{\varphi^{\prime}\left(g_{0}^{-1}(1-z)\right)}{g_{0}^{\prime}\left(g_{0}^{-1}(1-z)\right)} \cdot \frac{d z}{(t-z)^{1 / 2}}\right]
$$

## Putting

$$
\begin{aligned}
& g_{0}^{-1}(1-z)=u, \\
& g_{0}^{-1}(1-t)=y
\end{aligned}
$$

we have

$$
\frac{\left[g(y) g_{0}(y)\right]^{1 / 2}}{g_{0}^{\prime}(y)}=-\frac{\varphi(0)}{2 \pi\left[1-g_{0}(y)\right]^{1 / 2}}-\frac{1}{2 \pi} \int_{0}^{y} \frac{\varphi^{\prime}(u) d u}{\left[g_{0}(u)-g_{0}(y)\right]^{1 / 2}}
$$

or

$$
\left[g(y) g_{0}(y)\right]^{1 / 2}=-\frac{\varphi(0) g_{0}^{\prime}(y)}{2 \pi\left[1-g_{0}(y)\right]^{1 / 2}}-\frac{1}{2 \pi} g_{0}^{\prime}(y) \int_{0}^{y} \frac{\varphi^{\prime}(u) d u}{\left[g_{0}(u)-g_{0}(y)\right]^{1 / 2}} .
$$

Integrating from 0 to $\beta$, we get

$$
\begin{aligned}
\int_{0}^{\beta}\left[g(y) g_{0}(y)\right]^{1 / 2} d y= & -\frac{\varphi(0)}{2 \pi} \int_{0}^{\beta} \frac{g_{0}^{\prime}(y)}{\left[1-g_{0}(y)\right]^{1 / 2}} d y \\
& -\frac{1}{2 \pi} \int_{0}^{\beta} g_{0}^{\prime}(y)\left\{\int_{0}^{y} \frac{\varphi^{\prime}(u) d u}{\left[g_{0}(u)-g_{0}(y)\right]^{1 / 2}}\right\} d y \\
= & \frac{\varphi(0)}{\pi}\left[1-g_{0}(\beta)\right]^{1 / 2} \\
& -\frac{1}{2 \pi} \int_{0}^{\beta} \varphi^{\prime}(u)\left\{\int_{u}^{\beta} \frac{g^{\prime}(y) d y}{\left[g_{0}(u)-g_{0}(y)\right]^{1 / 2}}\right\} d u \\
= & \frac{\varphi(0)}{\pi}\left[1-g_{0}(\beta)\right]^{1 / 2} \\
& +\frac{1}{\pi} \int_{0}^{\beta} \varphi^{\prime}(u)\left[g_{0}(u)-g_{0}(\beta)\right]^{1 / 2} d u \\
= & -\frac{1}{\pi} \int_{0}^{\beta} \varphi_{0}(u)\left\{\frac{d}{d u}\left[g_{0}(u)-g_{0}(\beta)\right]^{1 / 2}\right\} d u
\end{aligned}
$$

$$
=\frac{1}{\pi} \int_{0}^{\beta} \varphi(u)\left\{-\frac{d}{d u}\left[g_{0}(u)-g_{0}(\beta)\right]^{1 / 2}\right\} d u
$$

We remark that $-d\left[g_{0}(u)-g_{0}(\beta)\right]^{1 / 2} / d u \geq 0$. Squaring and applying Schwarz's inequality, we have

$$
\begin{aligned}
& \left\{\int_{0}^{\beta}\left[g(y) g_{0}(y)\right]^{1 / 2} d y\right\}^{2} \\
& =\frac{1}{\pi^{2}}\left\{\int_{0}^{\beta} \varphi(u)\left(-\frac{d}{d u}\left[g_{0}(u)-g_{0}(\beta)\right]^{1 / 2}\right) d u\right\}^{2}, \\
& \int_{0}^{\beta} g(y) d y \cdot \int_{0}^{\beta} g_{0}(y) d y \geq \frac{a^{2}}{\pi^{2}}\left[1-g_{0}(\beta)\right]
\end{aligned}
$$

The equality sign holds when $g=g_{0}$ due to a converse part of the theorem on Schwarz's inequality. Then we have

$$
\begin{equation*}
\int_{0}^{\beta} g(y) d y \geq \frac{a^{2}}{\pi^{2}} \cdot \frac{1-g_{0}(\beta)}{\int_{0}^{\beta} g_{0}(y) d y} \tag{17}
\end{equation*}
$$

From (14), we can easily compute

$$
\begin{equation*}
1-g_{0}(\beta)=\left(\frac{e^{\pi \beta / \alpha}-1}{e^{\pi \beta / \alpha}+1}\right)^{2}=\frac{\pi^{2}}{4} k_{\alpha \beta}^{2} \tag{18}
\end{equation*}
$$

$$
\begin{equation*}
\int_{0}^{\beta} g_{0}(y) d y=\frac{2 \alpha}{\pi} \cdot \frac{e^{\pi \beta / \alpha}-1}{e^{\pi \beta / \alpha}+1}=k_{\alpha \beta} \alpha \tag{19}
\end{equation*}
$$

Combining (17), (18), and (19), we obtain

$$
\begin{equation*}
\int_{0}^{\beta} g(y) d y \geq k_{\alpha \beta} \frac{a^{2}}{4 \alpha} \tag{20}
\end{equation*}
$$

We remark that the equality sign in (20) holds for $g=g_{0}$. In fact, in this case, it can easily be proved that $a=2 \alpha$ by using the Weierstrass theory of extremal fields. From (19), we have

$$
\begin{equation*}
\int_{0}^{\beta} g_{0}(y) d y=\alpha k_{\alpha \beta}=k_{\alpha \beta} \frac{a^{2}}{4 \alpha} \tag{21}
\end{equation*}
$$

Combining the definition of area and (20), we obtain

$$
A=4 \alpha \int_{0}^{\beta} g(y) d y \geq k_{\alpha \beta} a^{2}
$$

By (21), we know that

$$
A=k_{\alpha \beta} a^{2}
$$

for $g=g_{0}$. This shows that our $k_{\alpha \beta}$ is the best constant.
3. Added in proof. In a course on Riemannian Geometry given at Syracuse University in 1949, Professor C. Loewner proved the inequality $A \geq 31 / 2 a 2 / 2$ for the case of $M_{t}^{2}$, the Riemannian manifold whose underlying topological space torus. The present investigation originates from this idea and has a similar method of treatment.

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# ON THE BARYCENTRIC HOMOMORPHISM IN A SINGULAR COMPLEX 

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## Introduction

0.1. Radó has introduced and studied the following approach to singular homology theory (see $[2 ; 3 ; 4]$ for details). With a general topological space $X$ associate a complex $R=R(X)$ in the following manner. For integers $p \geq 0$, let $v_{0}, \cdots, v_{p}$ be a sequence of $p+1$ points in Hilbert space $E_{\infty}$, which are not required to be distinct or linearly independent, and let $\left|v_{0}, \cdots, v_{p}\right|$ denote their convex hull. Suppose that $T$ is a continuous mapping from $\left|v_{0}, \cdots, v_{p}\right|$ into $X$. Then the sequence $v_{0}, \cdots, v_{p}$ jointly with $T$ determines a $p$-cell in $R$, which is denoted by $\left(v_{0}, \cdots, v_{p}, T\right)^{R}$. The free Abelian group $C_{p}^{R}$ generated by the $p$-cells in $R$ is termed the group of integral $p$-chains in $R$. For integers $p<0, C_{p}^{R}$ is defined to be the group consisting of the zero element alone. The boundary operator $\partial_{p}^{R}: C_{p}^{R} \rightarrow C_{p-1}^{R}$ is defined, in the usual manner, as the trivial homomorphism if $p \leq 0$, and by the relation

$$
\partial_{p}^{R}\left(v_{0}, \cdots, v_{p}, T\right)^{R}=\sum_{i=0}^{p}(-1)^{p}\left(v_{0}, \cdots, \hat{v}_{i}, \cdots, v_{p}, T\right)^{R}
$$

if $p>0$. Since $\partial_{p-1}^{R} \partial_{p}^{R}=0$, one introduces the subgroup $Z_{p}^{R}$ of $p$-cycles in $C_{p}^{R}$ and the subgroup $B_{p}^{R}$ of $p$-boundaries in $C_{p}^{R}$ in the customary way, and defines the quotient group of $Z_{p}^{R}$ with respect to $B_{p}^{R}$ to be the homology group $H_{p}^{R}$.
0.2. The approach to singular homology theory pursued by Radó differs from other approaches in that absolutely no identifications are made. Thus two $p$-cells $\left(v_{0}^{\prime}, \cdots, v_{p}^{\prime}, T^{\prime}\right)^{R}$ and $\left(v_{0}^{\prime \prime}, \cdots, v_{p}^{\prime \prime}, T^{\prime \prime}\right)^{R}$ are equal only if they are identical; that is, if $v_{i}^{\prime}=v_{i}^{\prime \prime}$ for $i=0, \cdots, p$ and $T^{\prime} \equiv T^{\prime \prime}$ on $\left|v_{0}^{\prime}, \cdots, v_{p}^{\prime}\right|$ $=\left|v_{0}^{\prime \prime}, \cdots, v_{p}^{\prime \prime}\right|$ In [3;4], Radó introduces a technique for making identifications in a general Mayer complex and applies his procedure to study identifications in $R$, particularly those which yield homology groups isomorphic to the $H_{p}^{R}$. It is a primary purpose of the present paper to pursue the matter further in
order to establish stronger results than those obtained by Radó.
The identification scheme of Radó for the complex $R$ is briefly described in §0.3 below; the reader should consult [3, §1] or [4, §5] for details.
0.3. Let $\left\{G_{p}\right\}$ be a collection of subgroups $G_{p}$ of the group $C_{p}^{R}$ of integral $p$-chains in $R$ such that $\partial_{p}^{R} G_{p} \subset G_{p-1}$ for every integer $p$; such a system is termed an identifier for $R$. Let $C_{p}^{m}$ be the quotient group of $C_{p}^{R}$ with respect to $G_{p}$, and denote that element of $C_{p}^{m}$ to which a chain $c_{p}^{R}$ in $C_{p}^{R}$ belongs by $\left\{c_{p}^{R}\right\}$. The restriction on the groups $G_{p}$ clearly implies that the element $\left\{\partial_{p}^{R} c_{p}^{R}\right\}$ in $C_{p-1}^{m}$ is independent of the choice of the representative $c_{p}^{R}$ of the element $\left\{c_{p}^{R}\right\}$ in $C_{p}^{m}$; thus one may define homomorphisms $\partial_{p}^{m}: C_{p}^{m} \longrightarrow C_{p-1}^{m}$ by the formula $\partial_{p}^{m}\left\{c_{p}^{R}\right\}=\left\{\partial_{p}^{R} c_{p}^{R}\right\}$. The resulting system of groups $C_{p}^{m}$ together with the operator $\partial_{p}^{m}$ constitutes a Mayer complex $m$ with homology groups $H_{p}^{m}$. Define a natural homomorphism $\pi_{p}: C_{p}^{R} \longrightarrow C_{p}^{m}$ by the formula $\pi_{p} c_{p}^{R}=\left\{c_{p}^{R}\right\}$. It is readily verified that $\pi_{p}$ is a chain mapping; hence it induces homomorphisms $\pi_{p}: H_{p}^{R} \rightarrow H_{p}^{m}$. If for every integer $p$ these homomorphisms are isomorphisms onto, then the identifier $\left\{G_{p}\right\}$ is termed unessential for $R$. liado notes that a necessary and sufficient condition in order that an identifier $G_{p}$ be unessential for $R$ is that every cycle $z_{p}^{R}$ in $G_{p}$ should be the boundary of some chain $c_{p+1}^{R}$ in $G_{p+1}$. (See [3, $\S \S 1.3,1.4,1.5]$ or [4, §5].)
0.4. One of the principal results in this paper may now be described. Let $\beta_{p}^{R}: C_{p}^{R} \rightarrow C_{p}^{R}$ be the barycentric homomorphism in $R$ (see [3, §3.1] or [4, §6]; also $\S 1.3$ ), and denote by $N\left(\beta_{p}^{R}\right)$ the nucleus of this homomorphism for every integer $p$.

Theorem. The system of nuclei $N\left(\beta_{p}^{R}\right)$ of the barycentric homomorphisms in in $R$ constitutes an unessential identifier for $R$ (see §3.2).

This result is combined with those of Rado in [3] to obtain stronger theorems concerning identifiers than any previously obtained. Since further definitions are necessary before these results can be described, the reader is requested to consult $\oint 3$ for the ir statements.
0.5. In the process of proving the theorem above, various results of independent interest have been attained. The reader is referred especially to $\S \S 1.6$, 1.7, 1.10, 2.2 for theorems which show the structural description of the barycentric homomorphism and of the barycentric homotopy operator.

## I. Further Relations in the Auxiliary Complex $K$

1.1. As in Radó [3;4], the auxiliary complex $K$ is the "formal complex", in the sense of [1], for the set $E_{\infty}$ of points in Hilbert space. For integers $p \geq 0$, $p$-cells in $K$ are ordered sequences ( $v_{0}, \cdots, v_{p}$ ) of $p+1$ points in $E_{\infty}$, which are not required to be distinct or linearly independent. These $p$-cells are taken as the base for a free Abelian group $C_{p}$, which is termed the group of finite integral $p$-chains in $K$. For $p<0$, the group $C_{p}$ is defined to be the group composed of the zero element alone. (See [3, §2.1] or [4, §6].)
1.2. In $K$ the following known homomorphisms will be used. (See [3, §2.2] or [4, §6].)
(i) For integers $j, p$ such that $0 \leq j \leq p, p>0$, the homomorphism

$$
j_{p}: C_{p} \rightarrow C_{p-1}
$$

is defined by the relation $j_{p}\left(v_{0}, \cdots, v_{p}\right)=(-1)^{j}\left(v_{0}, \cdots, \hat{v}_{j}, \cdots, v_{p}\right)$, where the symbol ${ }^{\wedge}$ is placed over the point $v_{j}$ to indicate that $v_{j}$ is to be deleted. For $j=p=0, j_{p}$ is defined to be the trivial homomorphism. A homomorphism differing from this one only by the absence of the factor $(-1)^{j}$ has been used by Radó in $[2, \S 2.6]$. The definition given above has been chosen because it permits simplifications in later definitions and formulas.
(ii) For integers $p>0$, the boundary operator

$$
\partial_{p}: C_{p} \longrightarrow C_{p-1}
$$

is defined by the formula

$$
\partial_{p}\left(v_{0}, \cdots, v_{p}\right)=\sum_{j=0}^{p}(-1)^{j}\left(v_{0}, \cdots, \hat{v}_{j}, \cdots, v_{p}\right)
$$

For integers $p \leq 0, \partial_{p}$ is defined to be the trivial homomorphism.
(iii) For integers $p \geq 0$ and an arbitrary point $v$ in $E_{\infty}$, the cone homomorphism $h_{p}^{v}: C_{p} \longrightarrow C_{p+1}$ is defined by the relation

$$
h_{p}^{v}\left(v_{0}, \cdots, v_{p}\right)=(-1)^{p+1}\left(v_{0}, \cdots, v_{p}, v\right) .
$$

For integers $p<0, h_{p}^{v}$ is defined to be the trivial homomorphism.
(iv) For integers $j, p$ such that $0 \leq j \leq p-1$, the transposition homomorphism $t_{p j}: C_{p} \longrightarrow C_{p}$ is defined by the relation

$$
t_{p j}\left(v_{0}, \cdots, v_{j}, v_{j+1}, \cdots, v_{p}\right)=\left(v_{0}, \cdots, v_{j+1}, v_{j}, \cdots, v_{p}\right)
$$

Observe that $t_{p j}\left(v_{0}, \cdots, v_{p}\right)=\left(v_{0}, \cdots, v_{p}\right)$ if and only if $v_{j}=v_{j+1}$.
(v) The barycentric homomorphism $\beta_{p}: C_{p} \longrightarrow C_{p}$ is defined as follows. For integers $p<0, \beta_{p}$ is the trivial homomorphism; for $p=0, \beta_{0}=1$; and for $p>0, \beta_{p}$ is defined by the recursion formula

$$
\beta_{p}\left(v_{0}, \cdots, v_{p}\right)=h_{p-1}^{b} \beta_{p-1} \partial_{p}\left(v_{0}, \cdots v_{p}\right)
$$

where $b$ is the barycenter of the points $v_{0}, \cdots, v_{p}$.
(vi) The barycentric homotopy operator $\rho_{p}$ used by Radó [1; 3, §2.2 (iv); $4, \S 6]$ will not be used in this paper. In its stead, a modification $\rho_{* p}$ is presently introduced, which has a simpler form, satisfies all the important identities which hold for the $\rho_{p}$, and has useful properties not possessed by $\rho_{p}$. The modified barycentric homotopy operator

$$
\rho_{* p}: C_{p} \longrightarrow C_{p+1}
$$

is defined as follows. For integers $p<0, \rho_{* p}$ is the trivial homomorphism; for $p=0, \rho_{* p}$ is defined by the relation

$$
\rho_{*_{0}}\left(v_{0}\right)=-h_{0}^{v_{0}}\left(v_{0}\right)=\left(v_{0}, v_{0}\right) ;
$$

and for $p>0, \rho_{* p}$ is defined by the recursion formula

$$
\rho_{* p}\left(v_{0}, \cdots, v_{p}\right)=-h_{p}^{b}\left[1+\rho_{* p-1} \partial_{p}\right]\left(v_{0}, \cdots, v_{p}\right),
$$

where $b$ is the barycenter of the points $v_{0}, \cdots, v_{p}$.
1.3. Amongst the preceding homomorphisms the following identities hold (see $[2, \S 2 ; 3, \S 2.3]):$

$$
\begin{array}{lr}
\partial_{p}=\sum_{j=0}^{p} j_{p} & (p \geq 0) ; \\
\partial_{p+1} h_{p}^{v}+h_{p-1}^{v} \partial_{p}=1 & (p>0) ; \\
\partial_{p} \beta_{p}=\beta_{p-1} \partial_{p} & (-\infty<p<+\infty) ; \\
\beta_{p} t_{p j}=-\beta_{p} & (0 \leq j \leq p-1) ; \\
\partial_{p+1} \rho_{* p}+\rho_{* p-1} \partial_{p}=\beta_{p}-1 & (0 \leq p<+\infty)
\end{array}
$$

Of these identities, only the last is new; it may be established by an inductive reasoning similar to that used to prove the corresponding identity for the conventional barycentric homotopy operator $\rho_{p}$.
1.4. For integers $k, p$ such that $0 \leq k \leq p$, the homomorphism

$$
k_{* p}: C_{p} \longrightarrow C_{p}
$$

is defined by the relation

$$
k_{* p}\left(v_{0}, \cdots, v_{p}\right)=(-1)^{p+k}\left(v_{0}, \cdots, \hat{v}_{k}, \cdots, v_{p}, v_{k}\right)
$$

and the homomorphism

$$
\gamma_{p}: C_{p} \longrightarrow C_{p}
$$

is defined by the formula $\gamma_{p}=\sum_{k=0}^{p} k_{* p}$. Obviously one has the identities

$$
\begin{aligned}
& k_{* p}\left(v_{0}, \cdots, v_{p}\right)=-k_{p+1} h_{p}^{v_{k}}\left(v_{0}, \cdots, v_{p}\right), p \geq 0 \\
& k_{* p}\left(v_{0}, \cdots, v_{p}\right)=h_{p-1}^{v_{k}} k_{p}\left(v_{0}, \cdots, v_{p}\right), p>0
\end{aligned}
$$

Now the reader will easily verify the relations

$$
\begin{gathered}
j_{p} k_{* p}=\left\{\begin{array}{cc}
(k-1)_{* p-1} j_{p} & , 0 \leq j<k \leq p, \\
k_{* p-1}(j+1)_{p} & , 0 \leq k \leq j<p, \\
k_{p} & , 0 \leq k \leq j=p ;
\end{array}\right. \\
k_{*-1} j_{p}=\left\{\begin{array}{cl}
(j-1)_{p} k_{* p} & , 0 \leq k<j \leq p, \\
j_{p}(k+1)_{* p} & , 0 \leq j \leq k<p
\end{array}\right.
\end{gathered}
$$

From these relations the following identity is readily established:

$$
\gamma_{p-1} \partial_{p}=\partial_{p}\left(\gamma_{p}-1\right)
$$

Using the identity, the reader will easily prove the following result.
Lemma. If $P(x)$ be any polynomial having integral coefficients, then

$$
P\left(\gamma_{p}-1\right) \partial_{p}=\partial_{p} P\left(\gamma_{p}-1\right)
$$

Explicitly, if $P(x)=\sum_{i=0}^{m} a_{i} x^{i}$, where the $a_{i}$ are integers, then

$$
\sum_{i=0}^{m} a_{i} \gamma_{p-1}^{i} \partial_{p}=\sum_{i=0}^{m} a_{i} \partial_{p}\left[\gamma_{p}^{i}-i \gamma_{p}^{i-1}+\cdots+(-1)^{i}\right]
$$

where $\gamma_{p}^{i}$ means that the homomorphism $\gamma_{p}$ is to be repeated itimes.
1.5. For integers $k, p$ such that $0 \leq k \leq p$, the homomorphism

$$
b_{p k}: C_{p} \longrightarrow C_{p+1}
$$

is defined by the relation

$$
\begin{aligned}
b_{p k}\left(v_{0}, \cdots, v_{p}\right)=(-1)^{k} & {\left[v_{0}, \cdots, v_{k}, b\left(v_{0}, \cdots, v_{k}\right),\right.} \\
& \left.b\left(v_{0}, \cdots, v_{k}, v_{k+1}\right), \cdots, b\left(v_{0}, \cdots, v_{k}, \cdots, v_{p}\right)\right]
\end{aligned}
$$

where $b\left(v_{0}, \cdots, v_{q}\right)$ is the barycenter of the points $v_{0}, \cdots, v_{q}$. Verification of the following simple relations is left to the reader:

$$
\begin{aligned}
& -h_{p}^{b\left(v_{0}, \cdots, v_{p}\right)}\left(v_{0}, \cdots, v_{p}\right)=b_{p p}\left(v_{0}, \cdots, v_{p}\right) ; \\
& -h_{p}^{b\left(v_{0}, \cdots, v_{p}\right)} b_{p-1 k}\left(v_{0}, \cdots, v_{p-1}\right)=b_{p k} h_{p-1}^{v_{p}}\left(v_{0}, \cdots, v_{p-1}\right) \\
& \\
& (0 \leq k \leq p-1) ; \\
& -h_{p}^{b\left(v_{0}, \cdots, v_{p}\right)} b_{p-1 k} j_{p}\left(v_{0}, \cdots, v_{p}\right)=b_{p k J * p}\left(v_{0}, \cdots, v_{p}\right) \\
& \quad(0 \leq k \leq p-1,0 \leq j \leq p) ; \\
& -h_{p}^{b\left(v_{0}, \cdots, v_{p}\right)} b_{p-1 k} \partial_{p}\left(v_{0}, \cdots, v_{p}\right)=b_{p k} \gamma_{p}\left(v_{0}, \cdots, v_{\mathrm{p}}\right)
\end{aligned}
$$

$$
(0 \leq k \leq p-1)
$$

If $P(x)$ be any polynomial having integral coefficients, then, for $0 \leq k \leq p-1$, we have

$$
-h_{p}^{b\left(v_{0}, \cdots, v_{p}\right)} b_{p-1 k} \partial_{p} P\left(\gamma_{p}\right)\left(v_{0}, \cdots, v_{p}\right)=b_{p k} \gamma_{p} P\left(\gamma_{p}\right)\left(v_{0}, \cdots, v_{p}\right) .
$$

1.6. For the homomorphisms $\beta_{p}$ and $\rho_{*}{ }^{\text {t }}$ the following structural descriptions are now obtained.

The orem. The following relations hold:

$$
\begin{aligned}
& \rho_{*_{0}}=b_{00} \\
& \rho_{* p}=b_{p p}+\sum_{j=1}^{p} b_{p p-j} \gamma_{p} \cdots\left(\gamma_{p}-j+1\right) \quad(p>0)
\end{aligned}
$$

Proof. It is sufficient to verify these formulas for a given $p-\operatorname{cell}\left(v_{0}, \cdots, v_{p}\right)$. For $p=0$, the formula $\rho_{* 0}\left(v_{0}\right)=b_{00}\left(v_{0}\right)$ is obvious from the definitions. So assume that

Using $\S 1.2, \S 1.4, \S 1.5$, and this assumption, and letting $b=b\left(v_{0}, \cdots, v_{p}\right)$, one obtains

$$
\begin{aligned}
& \rho_{* p}\left(v_{0}, \cdots, v_{0}\right) \\
& =-h_{p}^{b}\left(v_{0}, \cdots, v_{p}\right)-h_{p}^{b} \rho_{* p-1} \partial_{p}\left(v_{0}, \cdots, v_{p}\right) \\
& =b_{p p}\left(v_{0}, \ldots, v_{p}\right)-h_{p}^{b} b_{p-1 p-1} \partial_{p}\left(v_{0}, \ldots, v_{p}\right) \\
& -\sum_{j=1}^{p-1} h_{p}^{b} b_{p-1 p-1-j} \gamma_{p-1} \cdots\left(\gamma_{p-1}-j+1\right) \partial_{p}\left(v_{0}, \cdots, v_{p}\right) \\
& =b_{p p}\left(v_{0}, \cdots v_{p}\right)+b_{p p-1} \gamma_{p}\left(v_{0}, \cdots, v_{p}\right) \\
& -\sum_{j=1}^{p-1} h_{p}^{b} b_{p-1 p-1-j} \partial_{p}\left(\gamma_{p}-1\right) \cdots\left(\gamma_{p}-j\right)\left(v_{0}, \cdots, v_{p}\right) \\
& =b_{p p}\left(v_{0}, \cdots, v_{p}\right)+b_{p p-1} \gamma_{p}\left(v_{0}, \cdots, v_{p}\right) \\
& +\sum_{j=2}^{p} b_{p p-j} \gamma_{p}\left(\gamma_{p}-1\right) \cdots\left(\gamma_{p}-j+1\right)\left(v_{0}, \cdots, v_{p}\right) \\
& =b_{p p}+\sum_{j=1}^{p} b_{p p-j} \gamma_{p} \cdots\left(\gamma_{p}-j+1\right)\left(v_{0}, \cdots, v_{p}\right) .
\end{aligned}
$$

So the proof is complete by induction.
1.7. Theorem. The following relations hold:

$$
\begin{aligned}
& \beta_{0}=0_{1} b_{00} \\
& \beta_{p}=0_{p+1} b_{p 0} \gamma_{p}\left(\gamma_{p}-1\right) \cdots\left(\gamma_{p}-p+1\right), p>0 .
\end{aligned}
$$

The proof is similar to that for the theorm in the preceding section.
1.8. From these formulas for $\beta_{p}$ and $\rho_{* p}$ and the identities in $\S 1.3$, many further interesting relations may be obtained. For example, it is easy to establish the following results:

$$
\begin{array}{ll}
\beta_{p}=\left[\partial_{p+1}-(p+1)_{p+1}\right] \rho_{*_{p}} & (p \geq 0) ; \\
\partial_{p}=-p_{p} \rho_{* p-1} \partial_{p} & (p \geq 0) ; \\
\beta_{p}=(p+1)_{p+1}(p+2)_{p+2} \rho_{* p+1} \rho_{* p} & (p \geq 0) .
\end{array}
$$

These relations are not needed for the present purposes; they may be studied on a later occasion.

In order to clarify the structural descriptions for $\beta_{p}$ and $\rho_{* p}$ given in $\S \S 1.6$, 1.7 , it is convenient to introduce another homomorphism.
1.9. For integers $p \geq 0$, let $i_{0}$, $\cdot, i_{p}$ be any rearrangement of the sequence $0, \cdots, p$, and put $\epsilon_{i_{0}} \cdots i_{p}$ equal to +1 or to -1 according as $i_{0}, \cdots, i_{p}$ is obtained from $0, \cdots, p$ by an even or by an odd number of transpositions. With each rearrangement one associates a homomorphism

$$
\tau_{p}: C_{p} \longrightarrow C_{p}
$$

defined by the formula

$$
\tau_{p}\left(v_{0}, \cdots, v_{p}\right)=\epsilon_{i_{0}} \cdots i_{p}\left(v_{i_{0}}, \cdots, v_{i_{p}}\right)
$$

Sometimes, for clarity, the more explicit notation $\tau_{p}\left(i_{0}, \ldots, i_{p}\right)$ is used for this homomorphism. For integers $j$ such that $0 \leq j \leq p$, denote by $T_{p j}$ the class of all $\tau_{p}\left(i_{0}, \cdots, i_{p}\right)$ for which $i_{0}<\cdots<i_{j}-$ that is, for which $i_{0}, \cdots, i_{j}$ are in natural order. Obviously $T_{p p}$ consists of just one element, namely $\tau_{p}(0, \ldots, p)=1$; and $T_{p o}$ consists of the $\tau_{p}$ obtained by all possible rearrangements of $0, \cdots, p$. Moreover, $T_{p j-1} \supset T_{p j}$ for $1 \leq j \leq p$. Clearly the number of elements in the class $T_{p j}$ is $(p+1) p \cdots(j+2)$ for $0 \leq j \leq p-1$. For each integer $j$ in $0 \leq j \leq p$, define a homomorphism

$$
P_{p j}: C_{p} \longrightarrow C_{p}
$$

by the formula

$$
P_{p j}=\sum \tau_{p} \quad\left(\tau_{p} \in T_{p j}\right)
$$

Observe that $P_{p p}=1$. The reader will readily verify these identities:

$$
\begin{aligned}
& k_{* p} P_{p j}=P_{p j}, 0 \leq j<k \leq p \\
& \sum_{k=0}^{j} k_{*}, P_{p j}=P_{p j-1}, 0<j \leq p
\end{aligned}
$$

From these identities, the following result is established.
Lemma. The following relations hold:

$$
\begin{aligned}
P_{p p} & =1 \\
P_{p p-j} & =\gamma_{p}\left(y_{p}-1\right) \cdots\left(y_{p}-j+1\right), 1 \leq j \leq p
\end{aligned}
$$

Proof. That $P_{p p}=1$ was noted above. From the second relation above it follows that

$$
P_{p p-1}=\sum_{k=0}^{p} k_{* p} P_{p p}=\gamma_{p} P_{p p}=\gamma_{p}
$$

so the general formula is established for $j=1$. Now suppose that

$$
P_{p p-j+1}=\gamma_{p}\left(\gamma_{p}-1\right) \cdots\left(\gamma_{p}-j+2\right) \quad(2 \leq j \leq p)
$$

Using the preceding identities, one finds

$$
\begin{aligned}
\gamma_{p} P_{p p-j+1} & =\sum_{k=0}^{p} k_{* p} P_{p p-j+1} \\
& =\sum_{k=0}^{p-j+1} k_{* p} P_{p p-j+1}+\sum_{k=p-j+2}^{p} k_{* p} P_{p p-j+1} \\
& =P_{p p-j+(j-1) P_{p p-j+1}} \\
P_{p p-j} & =\left(\gamma_{p}-j+1\right) P_{p p-j+1}=\gamma_{p}\left(\gamma_{p}-1\right) \cdots\left(\gamma_{p}-j+1\right)
\end{aligned}
$$

Thus the lemma is established.
1.10. Combining the results of the preceding lemma with those in the theoems in §§l.6, 1.7, one obtains the following description for the homomorphisms $\beta_{p}$ and $\rho_{*}$.

Theorem. The following relations hold:

$$
\begin{array}{ll}
\beta_{p}=0_{p+1} b_{p 0} P_{p 0}=\sum_{\tau_{p} \in T_{p 0}} 0_{p+1} b_{p 0} \tau_{p} & (p \geq 0) ; \\
\rho_{* p}=\sum_{k=0}^{p} b_{p k} P_{p k}=\sum_{k=0}^{p} \sum_{\tau_{p}} \in T_{p k} b_{p k} \tau_{p} & (p \geq 0) .
\end{array}
$$

1.11. Let $v_{0}, \cdots, v_{p}(p \geq 0)$ be any sequence of $p+1$ points in $E_{\infty}$. In $\S \S 1.2,1.4,1.5,1.9$, homomorphisms $j_{p}, t_{p j}, k_{* p}, b_{p k}$, $\tau_{p}$, have been introduced which, when applied in any appropriate combination $h_{p}$ to the special chain $\left(v_{0}, \ldots, v_{p}\right)$, yield a special chain either of the form $+\left(y_{0}, \cdots, y_{q}\right)$ or of the form $-\left(y_{0}, \cdots, y_{q}\right)$. In the sequel, $\left[h_{p}\left(v_{0}, \cdots, v_{p}\right)\right]$ is defined to be the $p$-cell $\left(y_{0}, \ldots, y_{q}\right)$, and $\left|h_{p}\left(v_{0}, \cdots, v_{p}\right)\right|$ denotes its convex hull $\left|y_{0}, \cdots, y_{q}\right|$. For example,

$$
\begin{aligned}
& {\left[0_{p+1} b_{p 0} \tau_{p}\left(i_{0}, \cdots, i_{p}\right)\left(v_{0}, \cdots, v_{p}\right)\right] } \\
&=\left(b\left(v_{i_{0}}\right), b\left(v_{i_{0}}, v_{i_{1}}\right), \cdots, b\left(v_{i_{0}}, v_{i_{1}}, \cdots, v_{i_{p}}\right)\right)
\end{aligned}
$$

If for two sequences of points $u_{0}, \cdots, u_{p}$ and $v_{0}, \cdots, v_{p}$ it is true that

$$
\begin{aligned}
\left(b\left(u_{0}\right), b\left(u_{0}, u_{1}\right), \cdots, b\right. & \left.\left(u_{0}, u_{1}, \cdots, u_{p}\right)\right) \\
& =\left(b\left(v_{0}\right), b\left(v_{0}, v_{1}\right), \cdots, b\left(v_{0}, v_{1}, \cdots, v_{p}\right)\right)
\end{aligned}
$$

then clearly $u_{j}=v_{j}$ for $0 \leq j \leq p$. From the remarks in $\S 1.9$ and the preceding the orem, one thus obtains the following result.

Lemma. If the points $v_{0}, \cdots, v_{p}(p \geq 0)$ are distinct, then the chain $\beta_{p}\left(v_{0}, \cdots, v_{p}\right)$ contains $(p+1)$ ! terms; that is, for distinct elements $\tau_{p}^{\prime}$ and $\tau_{p}^{\prime \prime}$ in $T_{p 0}$, we have

$$
\left[0_{p+1} b_{p 0} \tau_{p}^{\prime}\left(v_{0}, \cdots, v_{p}\right)\right] \neq\left[0_{p+1} b_{p 0} \tau_{p}^{\prime \prime}\left(v_{0}, \cdots, v_{p}\right)\right]
$$

1.12. Lemma. Let $v_{0}, \cdots, v_{p}(p \geq 0)$ be any set of $p+1$ points in $E_{\infty}$,
not necessarily distinct or linearly independent. A necessary and sufficient condition that a point $v$ belong to the convex hull of the points
(i) $b\left(v_{0}\right), b\left(v_{0}, v_{1}\right), \cdots, b\left(v_{0}, v_{1}, \cdots, v_{p}\right)$
is that it possess a representation of the form
(ii) $v=\sum_{j=0}^{p} \mu_{j} v_{j} \quad\left(\sum_{j=0}^{p} \mu_{j}=1, \mu_{0} \geq \mu_{1} \geq \cdots \geq \mu_{p} \geq 0\right)$.

Proof. If $v$ belongs to the convex hull of the points (i), then it has a representation of the form
(iii) $v=\sum_{i=0}^{p} \lambda_{i} b\left(v_{0}, \cdots, v_{i}\right) \quad\left(\sum_{i=0}^{p} \lambda_{i}=1,0 \leq \lambda_{i}, 0 \leq i \leq p\right)$.

Thus

$$
v=\sum_{i=0}^{p} \lambda_{i} \sum_{j=0}^{i} \frac{v_{j}}{i+1}=\sum_{j=0}^{p} \sum_{i=j}^{p} \frac{\lambda_{i}}{i+1} v_{j},
$$

which gives a representation of form (ii) for $v$. Conversely, if $v$ has a representation of form (ii), put $\lambda_{i}=(i+1)\left(\mu_{i}-\mu_{i}+1\right)$ for $0 \leq i \leq p-1, \lambda_{p}=(p+1) \mu_{p}$. It follows at once that $v$ has a representation of form (iii), and hence belongs to the convex hull of the set of points (i).
1.13. For integers $p \geq 0$, if $u_{0}, \cdots, u_{p}$ is any sequence of $p+1$ points in $E_{\infty}$, then $\left|u_{0}, \cdots, u_{p}\right|$ will denote its convex hull. Let $k$ be any integer such that $0 \leq k \leq p$, and consider the sequence of $p+2$ points
(i) $u_{0}, \cdots u_{k}, b\left(u_{0}, \cdots u_{k}\right), \cdots, b\left(u_{0}, \cdots, u_{k}, \cdots, u_{p}\right)$,
that is (see $\S 1.5$ ), the sequence of points occurring in $b_{p k}\left(u_{0}, \cdots, u_{p}\right)$. Let
(ii) $w_{0}, \cdots, w_{p+1}$
be any rearrangement of the sequence of points (i). Designate by $x_{0}=w_{h_{0}}=u_{i_{0}}$ the first $u_{i}(0 \leq i \leq k)$ occurring in the sequence (ii). In general, let $x_{l}=w_{h_{l}}$ $=u_{i_{l}}(0 \leq l \leq k)$ be the $(l+1)$ st $u_{i}(0 \leq i \leq k)$ occurring in the sequence (ii), and put $x_{l}=u_{l}$ for $k+1 \leq l \leq p$ in case $k<p$. Now clearly $x_{0}, \cdots, x_{p}$ is a rearrangement of the sequence $u_{0}, \cdots, u_{p}$ in which the last $p-k$ elements are unaltered; the sequence (i) is a rearrangement of the sequence
(iii) $x_{0}, \cdots, x_{k}, b\left(x_{0}, \cdots, x_{k}\right), \cdots, b\left(x_{0}, \cdots, x_{k}, \cdots, x_{p}\right)$
in which the last $p+1-k$ elements are unaltered; and the sequence (ii) is a rearrangement of the sequence (iii) in which the points $x_{0}, \cdots, x_{k}$ appear in the same order as in (iii); that is, $x_{l}=w_{h_{l}}$ for $0 \leq l \leq k$, where $0 \leq h_{0}<h_{1}$ $<\ldots<h_{k} \leq p$. Now let $q$ be any integer such that $0 \leq q \leq p+1$. It will be shown that
(iv) $b\left(w_{0}, \cdots, w_{q}\right) \in\left|b\left(x_{0}\right), b\left(x_{0}, x_{1}\right), \cdots, b\left(x_{0}, x_{1}, \cdots, x_{p}\right)\right|$

$$
(0 \leq q \leq p+1)
$$

Case $q=0$. Then $b\left(w_{0}\right)=w_{0}$. If $w_{0}$ is one of the $u_{i}(0 \leq i \leq k)$, it follows by the choice above that $h_{0}=0$ and $w_{0}=x_{0}=b\left(x_{0}\right)$. If $w_{0}$ is not one of the $u_{i}(0 \leq i \leq k)$, there must be a $l \geq k$ such that $w_{0}=b\left(u_{0}, \cdots, u_{k}, \cdots, u_{l}\right)$ $=b\left(x_{0}, \cdots, x_{k}, \cdots, x_{l}\right)$. Thus relation (iv) is established when $q=0$.

General case. By a rearrangement, the points $w_{0}, \cdots, w_{q}$ may be ordered into two sets

$$
\begin{gathered}
w_{h_{0}}=x_{0}, \cdots, w_{h_{l}}=x_{l} \quad\left(0 \leq l \leq k, 0 \leq \dot{h}_{0}<\cdots<h_{l} \leq p\right), \\
\left\{\begin{array}{c}
w_{h_{l+1}}=b\left(u_{0}, \cdots, u_{k}, \cdots, u_{i_{l+1}}\right)=b\left(x_{0}, \cdots, x_{i_{l+1}}\right) \\
w_{h_{l+2}}=b\left(u_{0}, \cdots, u_{k}, \cdots, u_{i_{l+2}}\right)=b\left(x_{0}, \cdots, x_{i_{l+2}}\right) \\
\cdots \\
w_{h_{q}}=b\left(u_{0}, \cdots, u_{k}, \cdots, u_{i_{q}}\right)=b\left(x_{0}, \cdots, x_{i_{q}}\right) \\
\left(k \leq i_{l+1}<i_{l+2}<\cdots<i_{q} \leq p\right) .
\end{array}\right.
\end{gathered}
$$

The special cases which arise when one of these sets is missing are left to the reader. Now clearly

$$
\begin{aligned}
& b\left(w_{0}, \cdots, w_{q}\right)=b\left(w_{h_{0}}, \cdots, w_{h_{q}}\right) \\
& \quad=\sum_{j=0}^{l} \frac{1}{q+1}\left[1+\sum_{h=l+1}^{q} \frac{1}{i_{h}+1}\right] x_{j}+\sum_{j=l+1}^{i_{l+1}} \frac{1}{q+1} \sum_{h=l+1}^{q} \frac{1}{i_{h}+1} x_{j} \\
& \quad+\sum_{j=i_{l+1}+1}^{i_{l+2}} \frac{1}{q+1} \sum_{h=l+2}^{q} \frac{1}{i_{h}+1} x_{j}+\cdots+\sum_{j=i_{q-1}+1}^{i_{q}} \frac{1}{q+1} \frac{1}{i_{q}+1} x_{j} .
\end{aligned}
$$

In view of this equation and of the lemma in $\S 1.12$, the relation (iv) now follows.
1.14. From the facts presented above, the following result is presently established.

Lemma. Let $v_{0}, \cdots, v_{p}(p \leq 0)$ be any sequence of $p+1$ points in $E_{\infty}$. Fix $\tau_{p+1} \in T_{p+10}(0 \leq k \leq p), \tau_{p} \in T_{p k}$ (see §1.9). Then there exists $a$ $\tau_{p}^{\prime} \in T_{p 0}$ such that (see §1.11).

$$
\left|0_{p+2} b_{p+10} \tau_{p+1} b_{p k} \tau_{p}\left(v_{0}, \cdots, v_{p}\right)\right| \subset\left|0_{p+1} b_{p 0} \tau_{p}^{\prime}\left(v_{0}, \cdots, v_{p}\right)\right|
$$

Proof. Evidently $\left[\tau_{p}\left(v_{0}, \cdots, v_{p}\right)\right]=\left(v_{i_{0}}, \cdots, v_{i_{p}}\right)$, where $i_{0}, \cdots, i_{p}$ is a rearrangement of $0, \cdots, p$ such that $i_{0}<\cdots<i_{k}$. Put $u_{j}=v_{i_{j}}$ for $0 \leq j \leq p$, so that $\left[\tau_{p}\left(v_{0}, \cdots, v_{p}\right)\right]=\left(u_{0}, \cdots, u_{p}\right)$. Then

$$
\begin{aligned}
& {\left[b_{p k} \tau_{p}\left(v_{0}, \cdots, v_{p}\right)\right]} \\
& \quad=\left(u_{0}, \cdots, u_{k}, b\left(u_{0}, \cdots u_{k}\right), \cdots, b\left(u_{0}, \cdots, u_{k}, \cdots, u_{p}\right)\right)
\end{aligned}
$$

and $\left[\tau_{p+1} b_{p k} \tau_{p}\left(v_{0}, \cdots, v_{p}\right)\right]=\left(w_{0}, \cdots, w_{p+1}\right)$, where $w_{0}, \cdots, w_{p+1}$ is a rearrangement of

$$
u_{0}, \cdots, u_{k}, b\left(u_{0}, \cdots, u_{k}\right), \cdots, b\left(u_{0}, \cdots, u_{k}, \cdots, u_{p}\right)
$$

Finally,

$$
\begin{aligned}
& {\left[\begin{array}{lll}
0_{p+2} & b_{p+10} & \tau_{p+1}
\end{array} b_{p k} \tau_{p}\left(v_{0}, \cdots, v_{p}\right)\right]} \\
& \\
& \quad=\left[b\left(w_{0}\right), b\left(w_{0}, w_{1}\right), \cdots, b\left(w_{0}, w_{1}, \cdots, w_{p+1}\right)\right]
\end{aligned}
$$

The reasoning of $\oint 1.13$ shows that there is a rearrangement $x_{0}, \cdots, x_{p}$ of $u_{0}, \cdots, u_{p}$, and hence of $v_{0}, \cdots, v_{p}$, such that

$$
\begin{aligned}
\mid 0_{p+2} b_{p+10} \tau_{p+1} b_{p k} & \tau_{p}\left(v_{0}, \cdots, v_{p}\right) \mid \\
& \subset\left|b\left(x_{0}\right), b\left(x_{0}, x_{1}\right), \cdots, b\left(x_{0}, x_{1}, \cdots, x_{p}\right)\right| .
\end{aligned}
$$

Let $\tau_{p}^{\prime}$ be that element of $T_{p o}$ such that $\left[\tau_{p}^{\prime}\left(v_{0}, \cdots, v_{p}\right)\right]=\left(x_{0}, \cdots, x_{p}\right)$. Since

$$
\left[0_{p+1} b_{p 0} \tau_{p}^{\prime}\left(v_{0}, \cdots, v_{p}\right)\right]=\left(b\left(x_{0}\right), b\left(x_{0}, x_{1}\right), \cdots, b\left(x_{1}, x_{1}, \cdots, x_{p}\right)\right)
$$

the lemma is established.
1.15. If $c_{p}$ is a $p$-chain in $K$, and $A$ is a convex subset in $E_{\infty}$, then the in-
clusion $c_{p} \subset A$ will mean that either $c_{p}=0 \in C_{p}$ or else

$$
c_{p}=\sum_{j=1}^{n} m_{j}\left(v_{0 j}, \cdots, v_{p j}\right)
$$

where the $m_{j}$ are nonzero integers and $\left|v_{0 j}, \cdots, v_{p j}\right| \subset A$ for $1 \leq j \leq n$. One readily verifies the following inclusions (see [3, §2.4]):

$$
\begin{array}{rr}
j_{p}\left(v_{0}, \cdots, v_{p}\right) \subset\left|v_{0}, \cdots, v_{p}\right| & (0 \leq j \leq p), \\
\partial_{p}\left(v_{0}, \cdots, v_{p}\right) \subset\left|v_{0}, \cdots, v_{p}\right| & (p \geq 0), \\
\beta_{p}\left(v_{0}, \cdots, v_{p}\right) \subset\left|v_{0}, \cdots, v_{p}\right| & (p \geq 0), \\
\rho_{*_{p}}\left(v_{0}, \cdots, v_{p}\right) \subset\left|v_{0}, \cdots, v_{p}\right| & (p \geq 0), \\
t_{p j}\left(v_{0}, \cdots, v_{p}\right) \subset\left|v_{0}, \cdots, v_{p}\right| & (0 \leq j \leq p-1), \\
k_{* p}\left(v_{0}, \cdots, v_{p}\right) \subset\left|v_{0}, \cdots, v_{p}\right| & (0 \leq k \leq p), \\
\gamma_{p}\left(v_{0}, \cdots, v_{p}\right) \subset\left|v_{0}, \cdots, v_{p}\right| & (0 \leq k \leq p), \\
b_{p k}\left(v_{0}, \cdots, v_{p}\right) \subset\left|v_{0}, \cdots, v_{p}\right| & (0 \leq j \leq p), \\
\tau_{p}\left(v_{0}, \cdots, v_{p}\right) \subset\left|v_{0}, \cdots, v_{p}\right| & \left(T_{p 0}\right), \\
P_{p j}\left(v_{0}, \cdots, v_{p}\right) \subset\left|v_{0}, \cdots, v_{p}\right| & \\
\text { II. RELATIONS IN THE CoMPLEX R }=\text { R }(\mathrm{X}) . &
\end{array}
$$

2.1. If $A$ is a convex subset of $E_{\infty}$, then for integers $p \geq 0, C_{p}^{A}$ denotes that subgroup of $C_{p}$ generated by those $p$-cells $\left(v_{0}, \cdots, v_{p}\right)$ for which $\left|v_{0}, \cdots, v_{p}\right|$ $\subset A$; for $p<0$, we have $C_{p}^{A}=0 \in C_{p}$ (see $\S 1.1$ ). Suppose $T: A \longrightarrow X$ is a continuous mapping (see $\S 0.1$ ). For integers $p \geq 0$ define a homomorphism

$$
T_{p}: C_{p}^{A} \longrightarrow C_{p}^{R}
$$

by the relation $T_{p}\left(v_{0}, \cdots, v_{p}\right)=\left(v_{0}, \cdots, v_{p}, T\right)^{R}$ for $\left(v_{0}, \cdots, v_{p}\right) \in C_{p}^{A}$. For $p<0$, let $T_{p}$ be the trivial homomorphism. For chains $c_{p}$ in $C_{p}^{A}$ the notation $T_{p} c_{p}=\left(c_{p}, T\right)^{R}$ is used. In terms of this notation one finds the relation (see $\S 0.1): \partial_{p}^{R}\left(c_{p}, T\right)^{R}=\left(\partial_{p} c_{p}, T^{\prime}\right)^{R}$.

Now suppose that, for certain integers $p$,

$$
h_{p}: C_{p} \longrightarrow C_{q}
$$

is a homomorphism from the group $C_{p}$ of $p$-chains into the group $C_{q}$ of $q$-chains
in $K$ with the property that for all $p$-cells $\left(v_{0}, \cdots, v_{p}\right)$ in $K$ one has

$$
h_{p}\left(v_{0}, \cdots, v_{p}\right) \subset\left|v_{0}, \cdots, v_{p}\right|
$$

Then clearly one may define for these integers $p$ a homomorphism

$$
h_{p}^{R}: C_{p}^{R} \rightarrow C_{q}^{R}
$$

by the formula $h_{p}^{R}\left(v_{0}, \cdots, v_{p}, T\right)^{R}=\left(h_{p}\left(v_{0}, \cdots, v_{p}\right), T\right)^{R}$ in case $p \geq 0$, and one may make $h_{p}^{R}$ the trivial homomorphism if $p<0$. In view of the inclusions in §1.15, one observes that this definition creates the following homomorphisms in $R$ (see [3, §3.1]):

$$
\begin{aligned}
& j_{p}^{R}: C_{p}^{R} \rightarrow C_{p-1}^{R}(0 \leq j \leq p) ; \\
& \beta_{p}^{R}: C_{p}^{R} \rightarrow C_{p}^{R} \quad(-\infty<p<+\infty) ; \\
& \rho_{* p}^{R}: C_{p}^{R} \rightarrow C_{p}^{R} \rightarrow C_{p+1}^{R}(-\infty<p<+\infty) ;(p \geq 0) ; \\
& t_{p j}^{R}: C_{p}^{R} \rightarrow C_{p}^{R} \rightarrow C_{p-1}^{R}(0 \leq j \leq p-1) ; \tau_{p}^{R}: C_{p}^{R} \rightarrow C_{p}^{R} \quad(0 \leq k \leq p) ; \\
&\left.\tau_{p j}^{R}: C_{p}^{R} \rightarrow T_{p_{0}}\right) ;
\end{aligned}
$$

2.2. From the relations in $\S 1.3$, one derives the following (see [3, §3.1]):

$$
\begin{array}{lr}
\partial_{p}^{R} \beta_{p}^{R}=\beta_{p-1}^{R} \partial_{p}^{R} & (-\infty<p<+\infty) ; \\
\beta_{p}^{R} t_{p j}^{R}=-\beta_{p}^{R} & (0 \leq j \leq p-1) ; \\
\partial_{p+1}^{R} \rho_{* p}^{R}+\rho_{*_{p}-1}^{R} \partial_{p}^{R}=\beta_{p}^{R}-1 & (0 \leq p<+\infty) .
\end{array}
$$

The theorems in $\S \S 1.6,1,7$ give rise to these formulas for $\beta_{p}^{R}$ and $\rho_{* p}^{R}$ :

$$
\begin{aligned}
& \rho_{* 0}^{R}=b_{00}^{R}, \\
& \rho_{* p}^{R}=b_{p p}^{R}+\sum_{j=1}^{p} b_{p p-j}^{R} \gamma_{p}^{R}, \cdots,\left(\gamma_{p}^{R}-j+1\right)(p>0) ; \\
& \beta_{0}^{R}=0_{p}^{R} b_{00}^{R} ; \\
& \beta_{p}^{R}=0_{p+1}^{R} b_{p 0}^{R} \gamma_{p}^{R}\left(\gamma_{p}^{R}-1\right), \cdots,\left(\gamma_{p}^{R}-p+1\right)(p>0) .
\end{aligned}
$$

From the theorem in $\S 1.10$, one obtains the following description for $\beta_{p}^{R}$ and $\rho_{* p}^{R}$.

Theorem. The following relations hold:

$$
\begin{aligned}
& \beta_{p}^{R}=0_{p+1}^{R} b_{p 0}^{R}{\underset{p}{p 0}}_{R}^{R}=\sum_{\tau_{p} \in T_{p 0}} 0_{p+1}^{R} b_{p 0}^{R} \tau_{p}^{R}(p \geq 0) ; \\
& \rho_{* p}^{R}=\sum_{k=0}^{p} b_{p k}^{R} P_{p k}^{R}=\sum_{k=0}^{p} \sum_{\tau_{p}} T_{p k} b_{p k}^{R} \tau_{p}^{R}(p \geq 0) .
\end{aligned}
$$

2.3. The writer is indebted to T. Radó for suggestions which led to the results presently presented in $\S \delta 2.3-2.7,2.9,2.10,2,12$. The new facts contributed by this paper are contained in $\S \S 2.8,2.11,2.13$. For integers $p \geq 1$, any chain of the form $\left(1+t_{p i}^{R}\right)\left(v_{0}, \ldots, v_{p}, T\right)^{R}(0 \leq j \leq p-1)$ is termed an elementary $t$-chain in $R$ (see $[3, \S 3.2]$ or $[4, \S 7]$ ), and the subgroup of $C_{p}^{R}$ generated by these elementary $t$-chains is denoted by $T_{p}^{R}$. For $p<1, T_{p}^{R}$ is defined to be the subgroup of $C_{p}^{R}$ composed of the zero element alone.

Lemma. If $c_{p}^{R} \in T_{p}^{R}$, then
(i) $\partial_{p}^{R} c_{p}^{R} \in T_{p-1}^{R}$,
(ii) $\beta_{p}^{R} c_{p}^{R}=0$,
(iii) $\rho_{* p}^{R} c_{p}^{R} \in T_{p+1}^{R}$.

This lemma differs from that in Radó [3, §3.2], only by the fact that the barycentric homotopy operator $\rho_{p}^{R}$ has been replaced by the modified operator $\rho_{* p}^{R}$ (see §1.2). It may be established by the same reasoning as that employed by Radó.
2.4. For integers $p \geq 1$, any chain of the form

$$
\left(v_{0}, \cdots, v_{j}, v_{j+1}, \cdots, v_{p}, T\right)^{R}
$$

with $v_{j}=v_{j+1}$ for some $j$ such that $0 \leq j \leq p-1$ is called an elementary $d$-chain in $R$ (see $\left[3, \S 3.3\right.$ ] or $[4, \S 7]$ ), and the subgroup of $C_{p}^{R}$ generated by these elementary $d$-chains is denoted by $D_{p}^{R}$. For $p<1, D_{p}^{R}$ is defined to be that subgroup of $C_{p}^{R}$ composed of the zero element alone.

Lemma. If $c_{p}^{R} \in D_{p}^{R}$, then
(i) $\partial_{p}^{R} c_{p}^{R} \in D_{p-1}^{R}$,
(ii) $\beta_{p}^{R} c_{p}^{R}=0$,
(iii) $\rho_{* p}^{R} c_{p}^{R} \in D_{p+1}^{R}$.

This is the lemma in $[3, \S 3.3]$, except that the modified barycentric homotopy operator $\rho_{* p}^{R}$ is used in place of $\rho_{p}^{R}$; it is proved in the same way.
2.5. Lemma. Let $\left(v_{0}, \cdots, v_{p}, T\right)^{R}$ be any p-cell in $R(p \geq 1)$. Suppose that the sequence $w_{0}, \cdots, w_{p}$ is obtainable from the sequence $v_{0}, \cdots, v_{p}$ by $n$ transpositions. Then there is an element $t_{p}^{R}$ in $T_{p}^{R}$ such that

$$
\left(v_{0}, \cdots, v_{p}, T\right)^{R}=(-1)^{n}\left(w_{0}, \cdots, w_{p}, T\right)^{R}+t_{p}^{R}
$$

Proof. By assumption there exist $n+1$ sequences $v_{0} j, \cdots, v_{p j}$ for $0 \leq j \leq n$ where $v_{i 0}=v_{i}$ and $v_{i n}=w_{i}$ for $0 \leq i \leq p$ such that

$$
\left(v_{0 j}, \cdots, v_{p j}, T\right)^{R}=t_{p i_{j}}^{R}\left(v_{0 j-1}, \cdots, v_{p j-1}, T\right)^{R}
$$

for some integer $i_{j}$ satisfying $0 \leq i_{j} \leq p-1,1 \leq j \leq n$. Clearly

$$
\begin{aligned}
&\left(v_{0}, \cdots, v_{p}, T\right)^{R}=(-1)^{n}\left(w_{0}, \cdots, w_{p}, T\right)^{R} \\
&+\sum_{j=1}^{n}(-1)^{j-1}\left(1+t_{p i_{j}}^{R}\right)\left(v_{0 j-1}, \cdots, v_{p j-1}, T\right)^{R},
\end{aligned}
$$

and the lemma is established.
2.6. Lemma. Let $\left(v_{0}, \cdots, v_{p}, T\right)^{R}$ be any $p-c e l l$ in $R(p \geq 1)$, for which $v_{i}=v_{k}$ for some $i, k$ such that $0 \leq i \leq k \leq p$. Then there are elements $t_{p}^{R}$ in $T_{p}^{R}$ and $d_{p}^{R}$ in $D_{p}^{R}$ such that

$$
\left(v_{0}, \cdots, v_{p}, T\right)^{R}=t_{p}^{R}+d_{p}^{R}
$$

Moreover, $2\left(v_{0}, \cdots, v_{p}, T\right)^{R}$ is in $T_{p}^{R}$.
Proof. Since the sequence $v_{0}, \cdots, v_{i-1}, v_{k}, v_{i}, \cdots, v_{k-1}, v_{k+1}, \cdots, v_{p}$ is obtained from $v_{0}, \cdots, v_{i}, \cdots, v_{k}, \cdots, v_{p}$ by $k-i$ transpositions, and $v_{i}=v_{k}$ by assumption, if follows that

$$
(-1)^{k-i}\left(v_{0}, \cdots, v_{i-1}, v_{k}, v_{i}, \cdots, v_{k-1}, v_{k+1}, \cdots, v_{p}\right)^{R}
$$

is an element $d_{p}^{R}$ of $D_{p}^{R}$. Moreover, from the lemma in $\delta 2.5$ it follows that there is an element $t_{p}^{R}$ in $T_{p}^{R}$ such that $\left(v_{0}, \cdots, v_{p}, T\right)^{R}=d_{p}^{R}+t_{p}^{R}$, and the first part of the lemma is proven. Now the sequence $v_{0}, \cdots, v_{k}, \cdots, v_{i}, \cdots, v_{p}$ is obtained from $v_{0}, \cdots, v_{i}, \cdots, v_{k}, \cdots, v_{p}$ by $2(k-i)-1$ transpositions.

Again, from the lemma in $\delta 2.5$ it follows that there is an element $t_{p}^{R}$ in $T_{p}^{R}$ such that

$$
\left(v_{0}, \cdots, v_{i}, \cdots, v_{k}, \cdots, v_{p}, T\right)^{R}=-\left(v_{0}, \cdots, v_{k}, \cdots, v_{i}, \cdots, v_{p}, T\right)^{R}+t_{p}^{R}
$$

Since $v_{i}=v_{k}$, one obtains $2\left(v_{0}, \cdots, v_{p}, T\right)^{R}=t_{p}^{R}$; and the second part of the lemma is demonstrated.
2.7. For integers $p \geq 0$, a chain $c_{p}^{R}$ is termed an elementary $n$-chain in $R$ if it has the form

$$
c_{p}^{R}=\sum_{r=1}^{n} m_{r}\left(v_{0}, \cdots, v_{p}, T_{r}\right)^{R}
$$

where
(i) for $1 \leq r \leq n$, the $m_{r}$ are nonzero integers;
(ii) for $1 \leq r_{1}<r_{2} \leq n$, the transformations $T_{r_{1}}$ and $T_{r_{2}}$ are not identical on $\left|v_{0}, \cdots, v_{p}\right|$;
(iii) the points $v_{0}, \cdots, v_{p}$ are distinct. The $p$-cell $\left(v_{0}, \cdots, v_{p}\right)$ in $K$ (see $\S 1.11)$ is called the base for $c_{p}^{R}$, and the notation $c_{p}^{R}=c_{p}^{R}\left(v_{0}, \cdots, v_{p}\right)$ is used when it is desirable to display the base.
2.8. Lemma. Suppose that $c_{p}^{R}$ is an elementary $n$-chain in $R$ for which $\beta_{p}^{R} c_{p}^{R}=0$. Then $\beta_{p+1}^{R} \rho_{*_{p}}^{R} c_{p}^{R}=0$.

Proof. With the notation of $\S 2.7$, one finds (see $\S \S 2.1,2.2$ ).
(i) $\beta_{p}^{R} c_{p}^{R}=\sum_{\tau_{p} \in T_{p 0}} \sum_{r=1}^{n} m_{r}\left(0_{p+1} b_{p 0} \tau_{p}\left(v_{0}, \cdots, v_{p}\right), T_{r}\right)^{R}=0$;
(ii) $\beta_{p+1}^{R} \rho_{*_{p}}^{R} c_{p}^{R}=\sum_{\tau_{p+1} \in T_{p+1} 0} \sum_{k=0}^{p} \sum_{\tau_{p} \in T_{p k}} \sum_{r=1}^{n} m_{r}\left(0_{p+2} b_{p+10}\right.$

$$
\left.\tau_{p+1} b_{p k} \tau_{p}\left(v_{0}, \cdots, v_{p}\right), T_{r}\right)^{R}
$$

In view of $\S 2.7$ (iii), and $\S 1.11$, it follows from (i) that for each $\tau_{p}^{\prime} \in T_{p o}$, one has
(iii) $\sum_{r=1}^{n} m_{r}\left(0_{p+1} b_{p 0} \tau_{p}^{\prime}\left(v_{0}, \cdots, v_{p}\right), T_{r}\right)^{R}=0 \quad\left(\tau_{p}^{\prime} \in T_{p 0}\right)$,

Fix

$$
\tau_{p+1} \in T_{p+10}, \tau_{p} \in T_{p k} \quad(0 \leq k \leq p)
$$

From the lemma in $\oint 1.14$ follows the existence of a $\tau_{p}^{\prime} \in T_{p 0}$ such that
(iv) $\left|0_{p+2} b_{p+10} \tau_{p+1} b_{p k} \tau_{p}\left(v_{0}, \cdots, v_{p}\right)\right|$

$$
\subset\left|0_{p+1} b_{p 0} \tau_{p}^{\prime}\left(v_{0}, \cdots, v_{p}\right)\right|
$$

From (iii) and (iv) one concludes that for each

$$
\tau_{p+1} \in T_{p+10}, \tau_{p} \in T_{p k} \quad(0 \leq k \leq p)
$$

we have
(v) $\sum_{r=1}^{n} m_{r}\left(0_{p+2} b_{p+10} \tau_{p+1} b_{p k} \tau_{p}\left(v_{0}, \cdots, v_{p}\right), T\right)^{R}=0$.

In view of (ii) and (v) the lemma is now established.
2.9. For integers $p \geq 0$, the class $N_{p}^{R}$ is defined to be that subset of $C_{p}^{R}$ composed of the chain $0 \in C_{p}^{R}$ and of all $c_{p}^{R}$ having a representation of the form

$$
c_{p}^{R}=\sum_{s=1}^{n} c_{p s}^{R}\left(v_{0 s}, \cdots, v_{p s}\right)
$$

where
(i) for $1 \leq s \leq n$ the $c_{p s}^{R}\left(v_{0 s}, \cdots, v_{p s}\right)$ are elementary $n$-chains (see 2.7);
(ii) for $1 \leq s_{1}<s_{2} \leq n$, the point sets $v_{0 s_{1}}, \cdots, v_{p s_{1}}$ and $v_{0 s_{2}}, \cdots, v_{p s_{2}}$ are distinct. For $p<0$, the class $N_{p}^{R}$ consists of the chain $0 \in C_{p}^{R}$ alone. Each of the elementary $n$-chains $c_{p s}^{R}\left(v_{0 s}, \cdots, v_{p s}\right)(1 \leq s \leq n)$, is termed a $n$-composant of $c_{p}^{R}$. Observe that the sets $N_{p}^{R}$ are not generally subgroups of $C_{p}^{R}$.
2.10. Lemma. Let

$$
c_{p}^{R}=\sum_{s=1}^{n} c_{p s}^{R}\left(v_{0 s}, \cdots, v_{p s}\right)
$$

be any nonzero element in $N_{p}^{R}$. A necessary and sufficient condition in order that
$\beta_{p}^{R} c_{p}^{R}=0$ is that $\beta_{p}^{R} c_{p s}^{R}=0$ for every $n$-composant $c_{p s}^{R}(1 \leq s \leq n)$.
Proof. Trivially the condition suffices. It is presently shown to be necessary. With explicit notations (see $\S \delta 2.7,2.9$ ),

$$
\begin{aligned}
\beta_{p}^{R} c_{p}^{R} & =\sum_{s=1}^{n} \beta_{p}^{R} c_{p s}^{R}=\sum_{s=1}^{n} \sum_{r=1}^{n_{s}} m_{r s}\left(\beta_{p}\left(v_{0 s}, \cdots, v_{p s}\right) T_{r s}\right)^{R} \\
& =\sum_{s=1}^{n} \sum_{r=1}^{n_{s}} \sum_{\tau_{p}} \in_{T_{p 0}} m_{r s}\left(0_{p+1} b_{p 0} \tau_{p}\left(v_{0 s}, \cdots, v_{p s}\right), T_{r s}\right)^{R}=0 .
\end{aligned}
$$

In view of $\delta 2.9$ (ii) and of the remarks in $\delta 1.11$, it is clear (see $\oint 0.2$ ) that, for $1 \leq s \leq n$ we have

$$
\beta_{p}^{R} c_{p s}^{R}=\sum_{r=1}^{n_{s}} \sum_{\tau_{p} \in T_{p 0}} m_{r s}\left(0_{p+1} b_{p 0} \tau_{p}\left(v_{0 s}, \cdots, v_{p s}\right), T_{r s}\right)^{R}=0
$$

and hence the assertion in the lemma is verified.
2.11. Lemma. Let $c_{p}^{R}$ be any element in $N_{p}^{R}$ for which $\beta_{p}^{R} c_{p}^{R}=0$. Then

$$
\beta_{p+1}^{R} \rho_{* p}^{R} c_{p}^{R}=0
$$

This result is an immediate consequence of the lemmas in $\S \S 2.8,2.10$.
2.12. Lemma. Every chain $c_{p}^{R}$ has a representation of the form (see $\S \S 2.3$, 2.4, 2.9)

$$
c_{p}^{R}=t_{p}^{R}+d_{p}^{R}+n_{p}^{R} \quad\left(t_{p}^{R} \in T_{p}^{R}, d_{p}^{R} \in D_{p}^{R}, n_{p}^{R} \in N_{p}^{R}\right)
$$

Generally this representation is not unique.
Proof. The nonuniqueness of the representation will be evident from the proof of its existence which follows. For chains $c_{p}^{R}=0 \in C_{p}^{R}$, the result is trivial, so assume that $c_{p}^{R} \neq 0$. Then $c_{p}^{R}$ has a unique representation of the form
(i) $c_{p}^{R}=\sum_{j=1}^{n} m_{j}\left(v_{0 j}, \cdots, v_{p j}, T_{j}\right)^{R}$,
where the $m_{j}$ are nonzero integers and the $p$-cells $\left(v_{0 j_{1}}, \cdots, v_{p j_{1}}, T_{j_{1}}\right)^{R}$ and $\left(v_{0 j_{2}}, \cdots, v_{p j_{2}}, T_{j_{2}}\right)^{R}$ are distinct for $1 \leq j_{1} \leq j_{2} \leq n$. The proof is made by an induction on $n$. If $n=1$, then $c_{p}^{R}=m_{1}\left(v_{01}, \cdots, v_{p 1}, T_{1}\right)^{R}$. If, for some inte-
gers $i, k$ such that $0 \leq i<k \leq p$, one finds $v_{i_{1}}=v_{k 1}$, then the fact that $c_{p}^{R}$ has a representation of the prescribed form follows from the lemma in §2.6. On the other hand, if all the $v_{01}, \cdots, v_{p_{1}}$ are distinct, then $c_{p}^{R}$ is an elementary $n$-chain (see $\S 2.7$ ). Thus the lemma is established in case $n=1$. Suppose that the lemma is true for all chains $c_{p}^{R}$ having a representation of the form (i) with at most $n=N-1$ terms $(N>1)$. For chains $c_{p}^{R}$ whose representations (i) have $N$ terms it is convenient to consider several cases.

Case 1. Assume there is some term in the representation (i) of $c_{p}^{R}$ — without loss of generality one may assume it to be the first - for which there are integers $i, k$ such that $0 \leq i<k \leq p$ and $v_{i 1}=v_{k 1}$. By the lemma in $\delta 2.6$ there are elements $t_{p_{1}}^{R}$ in $T_{p}^{R}$ and $d_{p 1}^{R}$ in $D_{p}^{R}$ such that

$$
m_{1}\left(v_{01}, \cdots, v_{p 1}, T_{1}\right)^{R}=t_{p 1}^{R}+d_{p 1}^{R}
$$

By assumption there are elements $t_{p 2}^{R}$ in $T_{p}^{R}, d_{p 2}^{R}$ in $D_{p}^{R}$, and $n_{p}^{R}$ in $N_{p}^{R}$ such that

$$
\sum_{j=2}^{N} m_{j}\left(v_{0 j}, \cdots, v_{p j}, T_{j}\right)^{R}=t_{p 2}^{R}+d_{\dot{p} 2}^{R}+n_{p}^{R}
$$

Thus

$$
c_{p}^{R}=\left(t_{p 1}^{R}+t_{p 2}^{R}\right)+\left(d_{p 1}^{R}+d_{p 2}^{R}\right)+n_{p}^{R}
$$

and since $T_{p}^{R}$ and $D_{p}^{R}$ are subgroups of $C_{p}^{R}$, the existence of a representation of the prescribed form for $c_{p}^{R}$ follows in Case 1.

Case 2. Assume that for each $j(1 \leq j \leq N)$ the $v_{0 j}, \cdots, v_{p j}$ are distinct. By rearranging terms one may obtain from (i) a representation of the form

$$
\text { (ii) } c_{p}^{R}=\sum_{s=1}^{m} \sum_{r=1}^{n_{s}} m_{r s}\left(v_{0 s}, \cdots, v_{p s}, T_{r s}\right)^{R}, \sum_{s=1}^{m} n_{s}=N
$$

satisfying these conditions: none of the $m_{r s}$ is zero; for the same $s(1 \leq s \leq m)$, $1 \leq r_{1}<r_{2} \leq n_{s}$, the mappings $T_{r_{1} s}$ and $T_{r_{2} s}$ are not identical on $\mid v_{0 s}$, $\cdots, v_{p s} \mid ;$ for $1 \leq s_{1}<s_{2} \leq m$, the $p$-cells $\left(v_{0 s_{1}}, \cdots, v_{p s_{1}}\right)$ and ( $v_{0 s_{2}}$, $\cdots, v_{p s_{2}}$ ) are distinct in $K$ (see §l.1). Now for each $s(1 \leq s \leq m)$ clearly each of the chains

$$
c_{p s}^{R}=\sum_{j=1}^{n_{s}} m_{r s}\left(v_{0 s}, \cdots, v_{p s}, T_{r s}\right)^{R}
$$

is an elementary $n$-chain in $R$ (see $\oint 2.7$ ). The proof is carried forth by an inductive reasoning on $m$. If $m=1$ then $c_{p}^{R}$ is an elementary $n$-chain in $R$, and the representation (ii) already has the prescribed form. So assume that $c_{p}^{R}$, whose representation (i) has at most $N$ terms, has a representation of the prescribed form whenever its representation (ii) has at most $m=M-1$ terms ( $M>1$ ). Suppose now that $C_{p}^{R}$ is a chain whose representation (i) has $N$ terms while its representation (ii) has $M$ terms

$$
\sum_{s=1}^{M} n_{s}=N
$$

Subcase 2.1. Assume that for $1 \leq s_{1}<s_{2} \leq M$ the point sets $v_{0 s_{1}}, \cdots, v_{p s_{1}}$ and $v_{0 s_{2}}, \cdots, v_{p s_{2}}$ are distinct. From $\S 2.9$ it is clear that $c_{p}^{R}$ is itself an element in $N_{p}^{R}$ and representation (ii) has the prescribed form.

Subcase 2.2. Assume that there are distinct integers $s$ - with no loss of generality one may assume these to be $s=1$ and $s=2-$ such that the sets $v_{01}, \cdots, v_{p 1}$ and $v_{02}, \cdots, v_{p_{2}}$ are the same. It follows that the sequence $v_{02}, \cdots, v_{p 2}$ is obtainable from $v_{01}, \cdots, v_{p 1}$ by a positive number $l$ of transpositions. Hence by the lemma in $\oint 2.5$ there exists for each $r$ in $1 \leq r \leq n_{1}$ an element $t_{p r}^{R}$ in $T_{p}^{R}$ such that

$$
\left(v_{01}, \cdots, v_{p 1}, T_{r 1}\right)^{R}=(-1)^{l}\left(v_{02}, \cdots, v_{p 2}, T_{r_{1}}\right)^{R}+t_{p r}^{R} \quad\left(1 \leq r \leq n_{1}\right)
$$

Since $T_{p}^{R}$ is a subgroup of $C_{p}^{R}$, the chain

$$
\sum_{r=1}^{n_{1}} m_{r 1} t_{p r}^{R}
$$

is an element $t_{p *}^{R}$ in $T_{p}^{R}$. Consequently,

$$
\begin{aligned}
c_{p}^{R}=t_{p^{*}}^{R} & +\left[\sum_{r=1}^{n_{1}}(-1)^{l} m_{r 1}\left(v_{02}, \cdots, v_{p_{2}}, T_{r 1}\right)^{R}\right. \\
& \left.+\sum_{s=2}^{M} \sum_{r=1}^{n_{s}} m_{r s}\left(v_{0 s}, \cdots, v_{p s}, T_{r s}\right)^{R}\right] .
\end{aligned}
$$

Clearly the terms in square brackets may be rearranged into the form (ii) with an integer $m \leq M-1$, and the ir representation in form (i) has an integer $n \leq N$. By the inductive assumption there are elements $t_{p}^{R}$ in $T_{p}^{R}, d_{p}^{R}$, in $D_{p}^{R}$ and $n_{p}^{R}$ in
$N_{p}^{R}$ such that $c_{p}^{R}=\left(t_{p^{*}}^{R}+t_{p \#}^{R}\right)+d_{p}^{R}+n_{p}^{R}$, and the existence of a representation of the prescribed form for $c_{p}^{R}$ now follows in Case 2. Indeed, it is obvious in this case that $d_{p}^{R}=0 \in C_{p}^{R}$. So the lemma is completely established.
2.13. Lemma. If $c_{p}^{R}$ is any chain in $C_{p}^{R}$ for which $\beta_{p}^{R} c_{p}^{R}=0$, then

$$
\beta_{p+1}^{R} \rho_{* p}^{R} c_{p}^{R}=0
$$

The proof follows at once from the lemmas in $\oint \S 2.3,2.4,2.11,2.12$.

## Results

3.1. In [3, §4.1] (see also [4, §8]) Radó has established a lemma from which one derives the following statement by replacing the barycentric homotopy operator $\rho_{p}^{R}$ by the modified barycentric homotopy operator $\rho_{* p}^{R}$ (see $\S \S 1.2,2.1$ ).

Lemma. Let $\left\{G_{p}\right\}$ be an identifier for $R$ (see §0.3) such that the following conditions hold:
(i) $c_{p}^{R} \in G_{p}$ implies that $\beta_{p}^{R} c_{p}^{R}=0$;
(ii) $c_{p}^{R} \in G_{p}$ implies that $\rho_{*_{p}}^{R} c_{p}^{R} \in G_{p+1}$.

Then $\left\{G_{p}\right\}$ is unessential.
3.2. For each integer $p$ let $N\left(\beta_{p}^{R}\right)$ be the nucleus of the homomorphism $\beta_{p}^{R}: C_{p}^{R} \longrightarrow C_{p}^{R}$ (see $\delta 2.1$ ). Since $\beta_{p}^{R}$ is a chain mapping (see $\delta 2.2$ ) it is clear that the nuclei $N\left(\beta_{p}^{R}\right)$ constitute an identifier for $R$ (see $\oint 0.3$ ). Now in view of the lemma in $\$ 2.13$, conditions (i) and (ii) of the lemma above are clearly fulfilled for the identifier $\left\{N\left(\beta_{p}^{R}\right)\right\}$, and furthermore, this choice of an identifier yields the maximum amount of information that may be obtained from that lemma. Thus the $\left\{N\left(\beta_{p}^{R}\right)\right\}$ constitute an unessential identifier for $R$, and one of the main results is now established (see §0.4). It is summarized in the following statement.

The orem. The system of nuclei $N\left(\beta_{p}^{R}\right)$ of the barycentric homomorphisms $\beta_{p}^{R}: C_{p}^{R} \longrightarrow C_{p}^{R}$ constitutes an unessential identifier for $R$.
3.3. In order to compare this result with those in Radó [3; 4], first observe that it follows from the lemmas in $\oint \S 2.3,2.4$ that

$$
N\left(\beta_{p}^{R}\right) \supset T_{p}^{R}+D_{p}^{R} \quad(-\infty<p<+\infty)
$$

Moreover, since $C_{p}^{R}$ is a free group, it is clear that the division hull of $N\left(\beta_{p}^{R}\right)$
must be identical with the group $N\left(\beta_{p}^{R}\right)$. Thus the group $N\left(\beta_{p}^{R}\right)$ also contains the the division hull of the group $T_{p}^{R}+D_{p}^{R}$ for all integers $p$. An example is now given to show that the group $N\left(\beta_{p}^{R}\right)$ generally contains more.
3.4. Denote by $d_{0}, d_{1}, d$ the points $(1,0,0, \ldots),(0,1,0,0, \ldots),(1 / 2$, $1 / 2,0,0, \ldots)$ respectively, let $X$ be Euclidean $x$-space, and define transformations by the following relations:

$$
\left.\begin{array}{ll}
T_{1}: x=v_{0}-1 / 2 & \left(v \in\left|d_{0}, d_{1}\right|\right) ; \\
T_{2}: x= \begin{cases}0 & \left(v \in\left|d_{0}, d\right|\right) ; \\
v_{0}-1 / 2 & \left(v \in\left|d, d_{1}\right|\right) ;\end{cases} \\
T_{3}: x= \begin{cases}v_{0}-1 / 2 \\
0 & \left(v \in\left|d_{0}, d\right|\right) ;\end{cases} \\
T_{4}: x=0 & \left(v \in\left|d, d_{1}\right|\right) ;
\end{array}\right\}
$$

Clearly

$$
c_{1}^{R}=\left(d_{0}, d_{1}, T_{1}\right)^{R}-\left(d_{0}, d_{1}, T_{2}\right)^{R}-\left(d_{0}, d_{1}, T_{3}\right)^{R}+\left(d_{0}, d_{1}, T_{4}\right)^{R}
$$

belongs to $C_{1}^{R}$ and $\beta_{1}^{R} c_{1}^{R}=0$. Moreover, $c_{1}^{R}$ is an elementary $n$-chain (see §2.7). An elementary reasoning shows that it cannot belong to the division hull for the group $T_{1}^{R}+D_{1}^{R}$.
3.5. In order to describe the largest unessential identifier for $R$ obtained by Radó, a further definition is needed. For integers $p \geq 0$, let $\left(v_{0}, \cdots, v_{p}, T\right)^{R}$ bé any $p$-cell in $R$ (see $\S 0.1$ ). Let $w_{0}, \cdots, w_{p}$ be any set sequence of $p+1$ linearly independent points in $E_{\infty}$. Then there is a linear mapping

$$
\alpha:\left|w_{0}, \cdots, w_{p}\right| \rightarrow \mid v_{0}, \cdots, v_{p}
$$

such that $\alpha\left(w_{i}\right)=v_{i}$ for $0 \leq i \leq p$. The $p$-chain

$$
c_{p}^{R}=\left(v_{0}, \cdots, v_{p}, T\right)^{R}-\left(w_{0}, \cdots, w_{p}, T \alpha\right)^{R}
$$

is termed an elementary $a$-chain in $R$ (see [3, §3.4]), and the subgroup of $C_{p}^{R}$ generated by the elementary $a$-chains is denoted by $A_{p}^{R}$. For $p<0, A_{p}^{R}$ consists of the zero element alone. In [3, §3.4] Radó has a simple characterization for the group $A_{p}^{R}$ which he uses to define the group in [4, §7].
3.6. For each integer $p$, put $\Gamma_{p}^{R}=A_{p}^{R}+D_{p}^{R}+T_{p}^{R}$ (see $\S \S 2.3,2.4,3.5$ ), and let $\hat{\Gamma}_{p}^{R}$ denote the division hull of $\Gamma_{p}^{R}$. Then Radó shows that $\left\{\hat{\Gamma}_{p}^{R}\right\}$ is an
unessential identifier in $R$ (see $[3, \S 4.7]$ or $[4, \S 9]$ ), and this is his best result. If one sets $\Delta_{p}^{R}=A_{p}^{R}+N\left(\beta_{p}^{R}\right)$ (see $\S 3.2$ ) and lets $\hat{\Delta}_{p}^{R}$ denote the division hull of $\Delta_{p}^{R}$, then clearly $\Delta_{p}^{R} \supset \Gamma_{p}^{R}$, and hence $\hat{\Delta}_{p}^{R} \supset \hat{\Gamma}_{p}^{R}$. If one modifies the reasoning of Radó in [3, §4] by replacing the barycentric homotopy operator $\rho_{p}^{R}$ by the modified barycentric homotopy operator $\rho_{* p}^{R}$ (see $\delta 2.1$ ), one finds that $\hat{\Delta}_{p}^{R}$ is an unessential identifier for $R$. Thus one obtains the following result.

Theorem. If $\hat{\Delta}_{p}^{R}$ is the division hull of the group $A_{p}^{R}+N\left(\beta_{p}^{R}\right)$ then the system $\left\{\hat{\Delta}_{p}^{R}\right\}$ is an unessential identifier for $R$.

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## INCIDENCE RELATIONS IN MULTICOHERENT SPACES III

A. H. Stone

## 1. Introduction.

1.1. Preliminaries. The present paper is concerned with relations between systems of sets and their frontiers in a locally connected space $S$ of given degree of multicoherence, $r(S)$. The results are generalizations of those derived in [4] for the unicoherent case $[r(S)=0$ ], and of those in [5] for the case of two sets; the methods are those used in [5] and [6]. First we apply the "analytic" method (cf. [1; 2; 6]) to obtain a general 'addition theorem' for arbitrary sets with '"nearly disjoint'" frontiers (Theorem 1), which is shown to be "best possible" (Theorem 2), and to derive also relations between arbitrary systems of sets and their frontiers (Theorems 3 and 4). Next ( $\$ 4$ ) we consider a function of sets which measures (roughly speaking) the amount of disconnectedness of the frontiers of the components of the complementary set, and, after deriving some of its properties, use it to extend the Phragmén-Brouwer theorem to arbitrary sets (Theorem 6), and to obtain some related results. A modified 'addition theorem'" is then established (Theorem 9) which includes both Theorem 1 and Theorem 6 as special cases. Finally, we consider the incidences of sets with disjoint frontiers and subject to further restrictions (for example, that the sets be connected and have connected complements), showing that many problems of this type can be reduced to purely combinatorial problems in graph-theory.
1.2. Notations. We shall be concerned throughout with subsets of a fixed nonempty, connected, locally connected, completely normal ${ }^{1} T_{1}$ space, $S$. The notations are, in general, the same as in [4; 5; 6]; but the following items are repeated for the convenience of the reader.

The number of components, less one, of a set $E$, is denoted by $b_{0}(E)$; thus $b_{0}(0)=-1$. If the number of components of $E$ is infinite, we write $b_{0}(E)=\infty$, without distinction as to cardinality. The degree of multicoherence of $S$ is defined by $r(S)=\sup b_{0}(A \cap B)$, where $A$ and $B$ are closed connected sets such that $A \cup B=S$. It is known [5] that "closed"' can be replaced by "open' here.

If $A_{1}, A_{2}, \ldots, A_{n}$ are any $n$ sets (that is, subsets of $S$ ), and $J$ is any nonempty collection of distinct suffixes

[^2]$$
j_{1}, j_{2}, \cdots, j_{k} \quad\left(1 \leq j_{1}<j_{2}<\cdots<j_{k} \leq n\right),
$$
we write $A_{J}$ as an abbreviation for $A_{j_{1}} \cap A_{j_{2}} \cap \cdots \cap A_{j_{k}}$, and write
$$
\mathrm{U}\left\{A_{J}| | J \mid=k\right\} \text { as } X_{k}\left(A_{1}, A_{2}, \cdots, A_{n}\right),
$$
or simply as $X_{k}$. Thus
$$
\cup A_{j}=X_{1} \supset X_{2} \supset \cdots \supset X_{n}=\cap A_{j} .
$$

For convenience, we introduce the conventions $X_{0}=S$ and $X_{k}=0$ if $k>n$. We write

$$
\begin{equation*}
h\left(A_{1}, A_{2}, \cdots, A_{n}\right)=\sum b_{0}\left(X_{k}\right)-\sum b_{0}\left(A_{k}\right) \quad(1 \leq k \leq n), \tag{1}
\end{equation*}
$$

with the convention that in interpreting an equality or inequality involving $h\left(A_{1}, \cdots, A_{n}\right)$ in which $\sum b_{0}\left(A_{k}\right)=\infty$, we first transpose all negative terms. If the sets $A_{j}$ are all closed, or all open, or more generally have separated differences ${ }^{2}$, it is known [4, Th. 6b] that $h\left(A_{1}, \cdots, A_{n}\right) \geq 0$.

Again, following Eilenberg [1], we consider (continuous) mappings $f$ of subsets of $S$ into the circle $S^{1}$ of complex numbers of unit modulus, and write " $f \sim 1$ on $X$ " to mean that there exists a real (continuous) function $\phi$ on $X$ such that $f(x)=\exp [i \phi(x)]$ when $x \in X$. Mappings $f_{1}, f_{2}, \cdots, f_{m}$ of $X$ in $S^{1}$ are independent on $X$ if the only (positive or negative) integers $p_{1}, p_{2}, \cdots, p_{m}$, for which the product (in the sense of complex numbers)

$$
f_{1}^{p_{1}} f_{2}^{p_{2}} \cdots f_{m}^{p_{m}} \sim 1 \text { on } X,
$$

are $p_{1}=p_{2}=\cdots=p_{m}=0$. If $A_{1}, A_{2}, \cdots, A_{n}$ are closed sets whose union is $X$, the greatest number of mappings $f$ of $X$ in $S^{1}$ which are independent on $X$ and such that $f \sim 1$ on each $A_{j}$ (or $\infty$ if there is no such greatest number) is denoted by $p\left(A_{1}, A_{2}, \cdots, A_{n}\right)$. For fixed $X$ and $n$, we write

$$
\begin{equation*}
r_{n}(X)=\sup p\left(A_{1}, \cdots, A_{n}\right), \tag{2}
\end{equation*}
$$

the supremum being taken over all systems of $n$ closed sets $A_{1}, \cdots, A_{n}$ whose union is $X$. Clearly $0=r_{1}(X) \leq r_{2}(X) \leq \cdots$; it is known [1] that

$$
\sup _{n} r_{n}(X)=b_{1}(X)
$$

and $[1 ; 6]$ that $r_{2}(S)=r(S)$.

[^3]1.3. Some Lemmas. We shall require the following lemmas, some of which are known; the proofs of the rest are easy.
(1) If $A_{1}, A_{2}, \ldots, A_{n}$ have separated differences, then
(i) $\cup \operatorname{Fr}\left(A_{j}\right)=\cup \operatorname{Fr}\left(X_{j}\right)$;
(ii) $A_{j} \cap A_{k}$ and $\mathrm{Co}\left(A_{j} \cup A_{k}\right)$ are separated (l $\left.\leq j<k \leq n\right)$ if and only if $\mathrm{Cl}\left(X_{k}\right) \subset X_{j}$ and $X_{k} \subset \operatorname{Int}\left(X_{j}\right)$;
(iii) $\operatorname{Fr}\left(A_{j} \cap A_{k}\right) \cap \operatorname{Fr}\left(A_{j} \cup A_{k}\right)=0(1 \leq j<k \leq n)$ if and only if $X_{1}, X_{2}, \cdots, X_{n}$ have disjoint ${ }^{3}$ frontiers; that is, $\mathrm{Cl}\left(X_{k}\right) \subset \operatorname{Int}\left(X_{j}\right)$;
(iv) $A_{1}, A_{2}, \ldots, A_{n}$ are of finite incidence ${ }^{4}$ if and only if
$$
\sum b_{0}\left(X_{j}\right)<\infty .
$$
(2) If $A_{1}$ and $A_{2}$ are both open, or both closed, then $A_{1}-A_{2}$ and $A_{2}-A_{1}$ are separated; and further, $A_{1} \cap A_{2}$ and $\operatorname{Co}\left(A_{1} \cap A_{2}\right)$ are separated if and only if $\operatorname{Fr}\left(A_{1} \cap A_{2}\right) \cap \operatorname{Fr}\left(A_{1} \cup A_{2}\right)=0$. If $A_{1}$ and $A_{2}$ are open, this condition is equivalent to $\operatorname{Fr}\left(A_{1}\right) \cap \operatorname{Fr}\left(A_{2}\right) \cap \operatorname{Fr}\left(A_{1} \cap A_{2}\right)=0$.
(3) "Approximation lemma." If $A_{j}-A_{k}$ and $A_{k}-A_{j}$ are separated, and also $A_{j} \cap A_{k}$ and $\operatorname{Co}\left(A_{j} \cup A_{k}\right)$ are separated ( $1 \leq j<k \leq n$ ), then, given any open sets $W(J) \supset A_{J}$ (where $J$ runs over all nonempty sets of suffixes between 1 and $n$ ), there exist open sets $A_{j}^{*} \supset A_{j}$ such that, for any open sets $B_{j}$ satisfying $A_{j} \subset B_{j} \subset A_{j}^{*}$, we have $B_{J} \subset W(J)$ and
$$
\operatorname{Fr}\left(B_{j}\right) \cap \operatorname{Fr}\left(B_{k}\right) \cap \operatorname{Fr}\left(B_{j} \cap B_{k}\right)=0 \quad(1 \leq j<k \leq n)
$$

If further $\operatorname{Fr}\left(A_{j}\right) \cap \operatorname{Fr}\left(A_{k}\right) \cap \operatorname{Fr}\left(A_{j} \cup A_{k}\right)=0(j \neq k)$, the sets $A_{j}^{*}$ can be chosen so that the sets $B_{j}$ have disjoint frontiers.
(If $n=2$, this reduces to [5, Ths. 7 and 7a]; the general case follows by a straightforward induction over $n$ 。)
(4) If $A_{1}, A_{2}, \cdots, A_{n}$ are closed sets of finite incidence, then

$$
p\left(A_{1}, A_{2}, \cdots, A_{n}\right) \leq h\left(A_{1}, A_{2}, \cdots, A_{n}\right)
$$

if further no three of the sets $A_{j}$ have a common point (for example, if $n=2$ ), then $p=h$. (Cf. [6, §2.6].)

[^4](5) If $f$ maps $X$ in $S^{1}$, and $X$ is a finite union of disjoint closed sets on each of which $f \sim 1$, then $f \sim 1$ on $X$. (Trivial.)
(6) If $f$ maps $S$ in $S^{1}$, and $f \sim 1$ on a closed set $A \subset S$, then there exists an open set $V \supset A$ such that $f \sim 1$ on $V$. (Cf. [1, p.157; 6, §2.2 (2)].)
(7) If $f$ maps $A$ in $S^{1}$, and $f \sim 1$ on $\operatorname{Fr}(A)$, then $f$ may be extended to a mapping $f^{*}$ of $S$ in $S^{1}$ such that $f^{*} \sim 1$ on $\mathrm{Cl}(S-A)$.

For $f=\exp (i \phi)$ on $\operatorname{Fr}(A)$; since $\mathrm{Cl}(S-A)$ is normal, $\phi$ can be extended to a continuous real function $\phi^{*}$ on $\mathrm{Cl}(S-A)$; define $f^{*}=\exp \left(i \phi^{*}\right)$ on $\mathrm{Cl}(S-A)$, and $f^{*}=f$ elsewhere.
(8) If $A_{1}, A_{2}, \cdots, A_{n}$ are $n$ closed sets, and $1 \leq m \leq n$, then

$$
p\left(A_{1}, A_{2}, \cdots, A_{n}\right) \leq r_{m}\left(X_{1}\right)+r_{n+1-m}\left(X_{m}\right) \leq r_{m}\left(X_{1}\right)+b_{1}\left(X_{m}\right)
$$

For consider $N$ mappings $f_{1}, \cdots, f_{N}$ of $X_{1}\left(=\cup A_{j}\right)$ in $S^{1}$ which are independent on $X_{1}$ and satisfy $f_{k} \sim 1$ on $A_{j}(1 \leq k \leq N, 1 \leq j \leq n)$. We must prove

$$
N \leqq r_{m}\left(X_{1}\right)+r_{n+1-m}\left(X_{m}\right) .
$$

Let $s$ be the greatest number of mappings $f_{k}$ which are independent on $X_{m}$; since $X_{m} \subset A_{1} \cup A_{2} \cup \cdots \cup A_{n+1-m}$, clearly $s \leq r_{n+1-m}\left(X_{m}\right)$. We may suppose that the mappings $f_{k}$ are independent on $X_{m}$ for $N-s<k \leq N$, and then have, for each $k \leq N-s$, a relation of the form

$$
g_{k} \equiv f_{k}^{p_{k}} \prod_{t>N-s} f_{t}^{q_{k t} \sim 1}
$$

on $X_{m}$, where the exponents $p_{t}, q_{k t}$ are integers not all zero, so that clearly $p_{k} \neq 0$. It readily follows that the mappings $g_{k}(1 \leq k \leq N-s)$ of $X_{1}$ in $S^{1}$ are independent on $X_{1}$, and they clearly satisfy $g_{k} \sim 1$ on each $A_{j}$. Further, from (6) above, there exists an open set $V_{m} \supset X_{m}$ such that each $g_{k} \sim 1$ on $\mathrm{Cl}\left(V_{m}\right)$. Now $X_{m-1}-V_{m}$ is a finite union of disjoint closed sets of the form $A_{J}-V_{m}$ (where $|J|=k-1$ ), on each of which each $g_{k} \sim 1$ : hence, by (5), $g_{k} \sim 1$ on $X_{m-1}-V_{m}$, so that there exists an open set $V_{m-1} \supset X_{m-1}-V_{m}$ such that each $g_{k} \sim 1$ on $\mathrm{Cl}\left(V_{m-1}\right)$. Proceeding in this way, we obtain open sets

$$
V_{\lambda} \supset X_{\lambda}-\left(V_{\lambda+1} \cup V_{\lambda+2} \cup \cdots \cup V_{m}\right) \quad(1 \leq \lambda \leq m)
$$

such that each $g_{k} \sim 1$ on $\mathrm{Cl}\left(V_{\underline{\lambda}}\right)$. Since $\operatorname{UCl}\left(V_{\lambda}\right) \supset X_{1}$, the number $N-s$ of mappings $g_{k}$ is at most $p\left(\bar{V}_{1}, \bar{V}_{2}, \ldots, \bar{V}_{m}\right) \leq r_{m}\left(X_{1}\right)$, and the result follows.

As corollaries, we have:
(9) If, in the proof of (8), each of the mappings $f_{k} \sim 1$ on $X_{m}$, then $N \leq r_{m}\left(X_{1}\right)$.
(10) If no $m+1$ of the sets $A_{j}$ in (8) can have a common point (for example, if $m=n)$, then $p\left(A_{1}, A_{2}, \cdots, A_{n}\right) \leq r_{m}\left(X_{1}\right)$.

For in this case, $X_{m}$ falls into disjoint closed sets $A_{J}$, each contained in a single $A_{j}$; hence, from (5), each $f_{k} \sim 1$ on $X_{m}$.

## 2. An additional theorem.

2.1. Introduction. The last result, $1.3(10)$, combined with $1.3(4)$, gives another proof of the fact $[6$, Ths. 3 and 4 a$]$ that if $A_{1}, A_{2}, \cdots, A_{n}$ are closed sets which cover $S$, and no three of them have a common point, then

$$
h\left(A_{1}, \cdots, A_{n}\right) \leq r(S) .
$$

In the present section we shall obtain a considerable extension of this property (Theorem 1), and show that it is the "best possible" of its kind, incidentally obtaining a new characterization of $r(S)$ (Theorem 2).
2.2. Theorem 1. Let $A_{1}, A_{2}, \cdots, A_{n}$ be any subsets of $S$ having separated differences and such that $A_{j} \cap A_{k}$ and $\operatorname{Co}\left(A_{j} \cup A_{k}\right)$ are separated whenever $j \neq k .{ }^{5}$ Suppose that no point belongs to $A_{j}$ for more than $m$ distinct values of $j$, where $2 \leq m \leq n .{ }^{6}$ Then

$$
0 \leq h\left(A_{1}, A_{2}, \cdots, A_{n}\right) \leq(m-1) r(S) .
$$

Proof. Clearly we may assume that $r(S)$ and $b_{0}\left(A_{j}\right)$ are finite ( $1 \leq j \leq n$ ); from [5, Th. 9], the sets $A_{j}$ are then of finite incidence. Further, it will suffice to prove the theorem under the additional assumptions that the sets $A_{j}$ are closed and have disjoint frontiers. For if the theorem is known in this case, the method of "approximation" extends it first [applying the second part of $1.3(3)$ to the sets $\left.\mathrm{Co}\left(A_{j}\right)\right]$ to the case in which the sets $A_{j}$ are open and satisfy

[^5]$$
\operatorname{Fr}\left(A_{j}\right) \cap \operatorname{Fr}\left(A_{k}\right) \cap \operatorname{Fr}\left(A_{j} \cap A_{k}\right)=0
$$
and thence [by the first part of 1.3 (3)] to the general case; we omit the details, since the argument is a straightforward generalization of that in $[5, \S \S 7.4$ and 7.5] (cf. also $[6, \$ 4.4]$ ).

We write $X_{s}\left(A_{1}, A_{2}, \ldots, A_{t}\right)$ as $X_{s}^{t}(1 \leq s \leq t \leq n)$, and introduce the conventions $X_{s}^{t}=S$ if $1 \leq s \leq n<t$, or if $0=s<t$, and $X_{s}^{t}=0$ if $s>t$. Now (all the numbers involved being finite here) one readily verifies that
(1) $h\left(A_{1}, A_{2}, \cdots, A_{n}\right)=h\left(A_{1}, A_{2}, \cdots, A_{n-1}\right)$

$$
+\sum h\left(A_{n} \cap X_{s-1}^{n-1}, X_{s}^{n-1}\right) \quad(1 \leq s \leq n-1)
$$

and repeated application of this identity gives
(2) $h\left(A_{1}, A_{2}, \cdots, A_{n}\right)=\sum_{1}+\sum_{2}+\cdots+\sum_{n-1}$,
where

$$
\sum_{s}=\sum_{t} h\left(A_{t+1} \cap X_{s-1}^{t}, X_{s}^{t}\right) \quad(s \leq t \leq n-1) .
$$

We first show that
(3) $\sum_{s} \leq r(S)$ $(1 \leq s \leq n-1)$.

For, from 1.3 (4), we have

$$
\sum_{s}=\sum_{t p}\left(A_{t+1} \cap X_{s-1}^{t}, X_{s}^{t}\right) \quad(s \leq t \leq n-1)
$$

Let $f_{t j}$, where $j=1,2, \cdots, n_{t}$, be mappings of

$$
X_{s}^{t+1}=\left(A_{t+1} \cap X_{s-1}^{t}\right) \cup X_{s}^{t}
$$

in the unit circle such that
(i) $f_{t j} \sim 1$ on $A_{t+1} \cap X_{s-1}^{t}$,
(ii) $f_{t j} \sim 1$ on $X_{s}^{t}$,
(iii) for fixed $t$, these mappings are independent on $X_{s}^{t+1}$.

To prove (3), it suffices to show that the total number $\sum_{n_{t}}$
( $s \leq t \leq n-1$ ) of these mappings is at most $r(S)$.
We have

$$
\operatorname{Fr}\left[\operatorname{Cl}\left(S-X_{s}^{t+1}\right)\right] \subset \operatorname{Fr}\left(X_{s}^{t+1}\right) \subset \operatorname{Fr}\left(A_{t+1} \cap X_{s-1}^{t}\right) \cup \operatorname{Fr}\left(X_{s}^{t}\right)
$$

a union of two closed sets which are easily seen [from 1.3 (1)] to be disjoint. Hence [from $1.3(5)$ and $1.3(7)] f_{t j} \sim 1$ on $\operatorname{Fr}\left[\mathrm{Cl}\left(S-X_{s}^{t+1}\right)\right]$, and so $f_{t j}$ can be extended to a mapping, which we still denote by $f_{t j}$, of $S$ in the unit circle, in such a way that
(iv) $f_{t j} \sim 1$ on $\mathrm{Cl}\left(S-X_{s}^{t+1}\right)$.

We assert that the extended mappings $f_{t j}$ are all independent on $S$. For suppose not; then, for each $t$, there exists a mapping of the form

$$
g_{t}=\Pi_{j} f_{t j}^{p_{t j}} \quad\left(1 \leq j \leq n_{t}\right)
$$

where the exponents $p_{t j}$ are positive or negative integers, not all zero for all $t$, such that
(4) $g_{s} g_{s+1} \cdots g_{n-1} \sim 1$ on $S$.

From (ii), we have $g_{t} \sim 1$ on $X_{s}^{t}$; and so, if $t>s$, we have $g_{t} \sim 1$ on $X_{s}^{s+1}$. Thus (4) gives $g_{s} \sim 1$ on $X_{s}^{s+1}$; hence, from (iii), it follows that $g_{s}=1$, and all the exponents $p_{s j}$ are zero. A similar argument, with $s$ replaced by $s+1$, then proves $g_{s+1}=1$, and so on; finally all the exponents $p_{t j}$ must be zero, giving the desired contradiction.

Now write

$$
E_{k}=\operatorname{Cl}\left(X_{s}^{s+k-1}-X_{s}^{s+k-2}\right), \quad k=1,2, \cdots, n+2-s ;
$$

thus the sets $E_{k}$ are closed and cover $S$, and it is easy to see that no three of them have a common point. We shall show:
(5) $f_{t j} \sim 1$ on $E_{k}$.

In fact, if $k \leq t+1-s$, then $E_{k} \subset X_{s}^{s+k-1} \subset X_{s}^{t} ;$ if $k=t+2-s$, then $E_{k} \subset A_{t+1} \cap X_{s-1}^{t}$; and if $k \geq t+3-s$, then $E_{k} \subset \mathrm{Cl}\left(S-X_{s}^{t+1}\right)$; thus in each case (5) follows from (ii), (i), or (iv).

Thus the total number of mappings $f_{t j}$ is at most

$$
p\left(E_{1}, E_{2}, \cdots, E_{n+2-s}\right)
$$

but, by $1.3(10)$, this number is at most $r_{2}\left(\cup E_{k}\right)=r(S)$; thus (3) is established.

Now we further have $\Sigma_{s}=0$ if $s \geq m$, since the sets $A_{t+1} \cap X_{s-1}^{t}$ and $X_{s}^{t}$ are then disjoint (for $X_{s+1}=0$ ). Thus the theorem follows from (2) and (3).
2.3. Corollaries and Remarks. We make the following observations.
(1) For any two sets $A, B$, satisfying the hypotheses of Theorem 1 , we have $b_{0}(A)+b_{0}(B) \leq b_{0}(A \cup B)+b_{0}(A \cap B) \leq b_{0}(A)+b_{0}(B)+r(S)$. (this generalizes [5, Th. 9].)
(2) For any set $E$, we have

$$
b_{0}(\mathrm{Fr}(E)) \leq b_{0}(\bar{E})+b_{0}(\mathrm{Cl}(\mathrm{Co}(E))+r(S) .
$$

(this generalizes [4, §6.5].)
(3) In Theorem 1, the hypothesis that $A_{j} \cap A_{k}$ and $\operatorname{Co}\left(A_{j} \cup A_{k}\right)$ be separated $(j \neq k)$ may be omitted for each pair $j$, $k$ for which $A_{j} \subset A_{k}$; that is, it may be replaced by: For each $j, k(1 \leq j, k \leq n)$, either $A_{j} \subset A_{k}$, or $A_{j} \supset A_{k}$, or $A_{j} \cap A_{k}$ and $\operatorname{Co}\left(A_{j} \cup A_{k}\right)$ are separated. This is proved by noting that a more careful application of the approximation argument will still lead to closed sets with disjoint frontiers.
(4) Other results may be derived by observing that, under suitable conditions on the sets $A_{1}, \cdots, A_{n}$, further sums $\sum_{s}$ in $2.2(1)$ above will vanish. For example, Theorem 1 can be slightly sharpened as follows:

If $A_{1}, \ldots, A_{n}$ satisfy the hypotheses of Theorem 1 (as weakened in (3) above), and if they can be renumbered so that

$$
A_{\lambda+1} \subset A_{\lambda+2} \subset \cdots \subset A_{n},
$$

then

$$
h\left(A_{1}, \cdots, A_{n}\right) \leq \min (\lambda, m-1) r(S) .
$$

For the approximation argument enables us to assume, as before, that the sets $A_{1}, \cdots, A_{n}$ are closed and have disjoint frontiers. In 2.2 (2) we easily verify that now $X_{s}^{t} \subset A_{t+1} \cap X_{s-1}^{t}$ whenever

$$
\lambda+1 \leq s \leq t \leq n-1
$$

hence $\sum_{s}=0$ whenever $s>\lambda$.
(5) A further slight sharpening of Theorem 1 is implied by the following result.

If the sets $A_{1}, \cdots, A_{n}$ have separated differences, and if $A_{n}(s a y)$ is either disjoint from, or contains, or is contained in, each other set, then

$$
h\left(A_{1}, \ldots, A_{n-1}, A_{n}\right)=h\left(A_{1}, \ldots, A_{n-1}\right)
$$

We may assume that $A_{n}$ is disjoint from $A_{1}, \cdots, A_{k}$, contains

$$
A_{k+1}, \cdots, A_{l}
$$

and is contained in $A_{l+1}, \cdots, A_{n-1}$ (where $\left.0 \leq k \leq l \leq n-1\right)$. It is easy to see that we may take $A_{1}, \cdots, A_{n}$ to be of finite incidence, and then, by $2.2(1)$, have only to prove that

$$
h\left(A_{n} \cap X_{s}^{n-1}, X_{s}^{n-1}\right)=0 \quad(1 \leq s \leq n-1)
$$

If $s<n-l$, then $A_{n} \subset X_{s}^{n-1}$, and the result is trivial. If $s \geq n-l$, write

$$
Y_{p}=X_{p}\left(A_{1}, \cdots, A_{k}, A_{l+1}, \cdots, A_{n-1}\right)
$$

and

$$
Z_{q}=Z_{q}\left(A_{k+1}, \cdots, A_{l}, A_{l+1}, \cdots, A_{n-1}\right)
$$

it is easily verified that $X_{s}^{n-1}=Y_{s} \cup Z_{s}$ and that $Y_{s} \subset \operatorname{Co}\left(A_{n}\right)$ and $Z_{s} \subset A_{n}$, from which again the result follows.
(6) Finally, as a corollary from (4), we have the following extension of (1):

If $B_{1}, \ldots, B_{p}, C_{1}, \cdots, C_{q}$ are arbitrary sets such that $B_{j}-C_{k}$ and $C_{k}-B_{j}$ are separated, and $B_{j} \cap C_{k}$ and $\operatorname{Co}\left(B_{j} \cup C_{k}\right)$ are separated, whenever $1 \leq j \leq p, 1 \leq k \leq q$, then

$$
\begin{aligned}
& h\left(B_{1}, \cdots, B_{p}\right)+h\left(C_{1}, \cdots, C_{q}\right) \leq h\left(B_{1}, \cdots, B_{p}, C_{1}, \cdots, C_{q}\right) \\
& \quad \leq h\left(B_{1}, \cdots, B_{p}\right)+h\left(C_{1}, \cdots, C_{q}\right)+\min (p, q, m-1) r(S)
\end{aligned}
$$

where $m$ is the greatest number of the $p+q$ sets $B_{1}, \ldots, C_{q}$ which have a common point.

This follows on application of (4) and Theorem 1 to the $p+q$ sets $X_{j}\left(B_{1}, \cdots, B_{p}\right), X_{k}\left(C_{1}, \cdots, C_{q}\right)$.
2.4. Converse. The converse of Theorem 1 holds in the following rather strong form, which represents an extension to any number of sets of the defining property of $r(S)$.

Theorem 2. Let integers $m$, $n$ be given, where $2 \leq m \leq n$. Let $A_{1}, \cdots, A_{n}$ be any $n$ closed connected sets, no $m+1$ of which have a common point, such that $\operatorname{Fr}\left(A_{j}\right) \cap \operatorname{Fr}\left(A_{k}\right)=0$ whenever $j \neq k$, and such that $A_{j} \cup A_{k}=S$ whenever $1 \leq j<k \leq m .^{7}$ Then

$$
\sup b_{0}\left(X_{m}\right)=(m-1) r(S)+n-m
$$

In this statement, the word "closed" may be replaced by "open".
To show that
(1) $b_{0}\left(X_{m}\right) \leqq(m-1) r(S)+n-m$,
we clearly may assume $X_{m} \neq 0$; then $b_{0}\left(X_{s}\right) \geq 0$ if $s \leq m$, and

$$
b_{0}\left(X_{s}\right)=-1
$$

for $m<s \leq n$, so that ( 1 ) is a trivial consequence of Theorem 1 .
To complete the proof, let $N$ be any integer such that

$$
0 \leq N \leq(m-1) r(S)
$$

We first construct $m$ closed connected sets $B_{1}, B_{2}, \cdots, B_{m}$, such that
(2) $B_{j} \cup B_{k}=S(1 \leq j<k \leq m) \quad$ and $\quad b_{0}\left(\cap B_{j}\right) \geq N$.

If $r(S)=\infty$, this is trivial (take all but two of the sets $B_{j}$ to be $S$ ), so we may assume $r(S)<\infty$. From [6, §4.1], there exists a finite covering of $S$ by closed connected sets $E_{1}, E_{2}, \ldots, E_{M}$, no three of which háve a common point, whose nerve $G$ satisfies $r(G)=r(S)=r$, say, and such that $G$ is arbitrarily often "dispersed"; this implies [6, §3.4(7)] that $G$ is obtainable from a graph $H$ by subdividing each arc $l_{\lambda}$ of $H$ which belongs to a simple closed curve in $H$, into at least $2 m+2$ subarcs by extra vertices of order 2 . We can select ${ }^{8} r$ such (disjoint, open)

[^6]${ }^{8}$ See, for example, the argument proving $[6, \S 4.1$ (3)].
arcs $l_{\lambda}$ in $H$, say $l_{1}, l_{2}, \cdots, l_{r}$, whose removal does not disconnect $H$; let $l_{\lambda}$ (where $1 \leq \lambda \leq r$ ) contain the consecutive vertices $p_{\lambda, 0}, p_{\lambda, 1}$, $p_{\lambda, 2}, \cdots, p_{\lambda, 2 m}$ of order 2 in $G$. Denote by $E_{\lambda, j}$ the set $E_{k}$ which corresponds to $p_{\lambda, j}$; thus, if $1 \leq \lambda \leq r$ and $1 \leq j \leq 2 m-1$, each $E_{\lambda, j}$ meets two and only two other sets $E_{k}$, namely $E_{\lambda, j-1}$ and $E_{\lambda, j+1}$. Define $B_{q}$, where $l \leq q \leq m$, to be the union of all the sets $E_{k}$ except
$$
E_{1,2 q-1}, E_{2,2 q-1}, \cdots, E_{r, 2 q-1}
$$

Then $B_{q}$ is closed, and is easily seen to be connected (cf. [6, Th. 1]). Further, since $\operatorname{Co}\left(B_{q}\right) \subset U_{\lambda} E_{\lambda, 2 q-1}$, we have $\operatorname{Co}\left(B_{q}\right) \cap \operatorname{Co}\left(B_{s}\right)=0$ if $q \neq s$, so that $B_{q} \cup B_{s}=S$. On the other hand, let $D$ be the union of those sets $E_{k}$ which are not of the form $E_{\lambda_{j} j}(1 \leq \lambda \leq r, \quad 1 \leq j$ $\leq 2 m-1$ ); then

$$
\cap B_{q} \subset D \cup \cup E_{\lambda, 2 h} \quad(1 \leq \lambda \leq r, 1 \leq h \leq m-1)
$$

a union of $1+(m-1) r$ disjoint closed sets, each of which it meets; thus $b_{0}\left(\cap B_{q}\right) \geq(m-1) r \geq N$.

There exist (cf. 1.3 (3) and $[6, \S 6.1]$ ) connected open sets $C_{q} \supset B_{q}$ whose closures $A_{q}$ have the same incidences as the sets $B_{q}$; then

$$
\operatorname{Fr}\left(A_{j}\right) \cap \operatorname{Fr}\left(A_{k}\right) \subset \operatorname{Fr}\left(C_{j}\right) \cap \operatorname{Fr}\left(C_{k}\right) \subset \operatorname{Co}\left(C_{j} \cup C_{k}\right)=0
$$

whenever $j \neq k$, and moreover we have $A_{j} \cup A_{k}=S(1 \leq j<k \leq m)$ and $b_{0}\left(\cap A_{j}\right) \geq N$.

If $n=m$, the theorem is thus established. If $n>m$, we note that the open set Int $\left[X_{m-1}\left(A_{1}, \cdots, A_{m}\right)\right]-X_{m}\left(A_{1}, \ldots, A_{m}\right)$ is nonempty, from $1.3(1)$, and take $A_{m+1}, \cdots, A_{n}$ to be $n-m$ distinct points in it; clearly

$$
b_{0}\left[X_{m}\left(A_{1}, \cdots, A_{m}, \cdots, A_{n}\right)\right] \geq N+n-m
$$

and the proof is complete.
The modifications required to produce open sets $A_{j}$ with similar properties are obvious.

## 3. Index inequalities for arbitrary sets.

3.1. An Inequality. Let $E_{1}, E_{2}, \cdots, E_{n}$ be arbitrary subsets of $S$. As
in [4, §7], we write

$$
A_{j}=\mathrm{Cl}\left(E_{j}\right), B_{j}=\mathrm{Cl}\left(S-E_{j}\right), P_{j}=X_{j}\left(A_{1}, \cdots, A_{n}\right), Q_{j}=X_{j}\left(B_{1}, \cdots, B_{n}\right) .
$$

An argument entirely analogous to that in [4, §7], based on 2.3 (1) and (2), gives:
Theorem 3. We have

$$
\begin{aligned}
h\left[\operatorname{Fr}\left(E_{1}\right), \cdots, \operatorname{Fr}\left(E_{n}\right)\right]-n r(S) & \leq h\left(A_{1}, \cdots, A_{n}\right)+h\left(B_{1}, \cdots, B_{n}\right) \\
+ & h\left(P_{1} \cap Q_{n}, P_{2} \cap Q_{n-1}, \cdots, P_{n} \cap Q_{1}\right) \\
& \leq h\left[\operatorname{Fr}\left(E_{1}\right), \cdots, \operatorname{Fr}\left(E_{n}\right)\right]+n r(S)
\end{aligned}
$$

Corollary. We have

$$
h\left(\bar{E}_{1}, \bar{E}_{2}, \cdots, \bar{E}_{n}\right) \leq h\left[\operatorname{Fr}\left(E_{1}\right), \operatorname{Fr}\left(E_{2}\right), \cdots, \operatorname{Fr}\left(E_{n}\right)\right]+n r(S)
$$

3.2. The Case $m=2$. It is easy to see that the inequalities in Theorem 3 are "best possible"; however, Theorem 1 suggests that in the Corollary the term $n r(S)$ could be replaced by $(n-1) r(S)$, or more generally by $(m-1) r(S)$, where no $m+1$ of the sets $\mathrm{Cl}\left(E_{j}\right)$ have a common point. I have been able to prove this only in the case $m=2$ :

Theorem 4. If $E_{1}, E_{2}, \ldots, E_{n}$ are arbitrary subsets of $S$, no three of whose closures have a common point, then

$$
h\left(\bar{E}_{1}, \bar{E}_{2}, \ldots, \bar{E}_{n}\right) \leq h\left[\operatorname{Fr}\left(E_{1}\right), \operatorname{Fr}\left(E_{2}\right), \ldots, \operatorname{Fr}\left(E_{n}\right)\right]+r(S)
$$

Proof. We can assume that $r(S)$ is finite, and that the systems of sets $\left[\mathrm{Cl}\left(E_{1}\right), \ldots, \mathrm{Cl}\left(E_{n}\right)\right]$ and $\left[\mathrm{Fr}\left(E_{1}\right), \ldots, \operatorname{Fr}\left(E_{n}\right)\right]$ are both of finite incidence, since otherwise (in view of the convention regarding infinite terms in the $h$ function; see 1.2) Theorem 4 asserts no more than Theorem 3, Corollary. Hence, in view of 1.3(4), Theorem 4 will follow [if we take $A_{j}=\mathrm{Cl}\left(E_{j}\right)$ and $F_{j}=\operatorname{Fr}\left(E_{j}\right)$ ] from:

Theorem 4a. Let $A_{1}, A_{2}, \ldots, A_{n}, F_{1}, F_{2}, \ldots, F_{n}$ be any closed sets such that $A_{j} \supset F_{j}$ and $\cup F_{j} \supset \cup \operatorname{Fr}\left(A_{j}\right)$. Then

$$
p\left(A_{1}, A_{2}, \cdots, A_{n}\right) \leq p\left(F_{1}, F_{2}, \cdots, F_{n}\right)+r(S) .
$$

3.3. Proof of Theorem 4a. Let $f_{1}, f_{2}, \cdots, f_{N}$ be $N$ independent mappings of $\cup A_{j}$ in the unit circle such that each $f_{k} \sim 1$ on each $A_{j}$; we must prove that $N \leq p\left(F_{1}, \cdots, F_{n}\right)+r(S)$. Let $s$ be the greatest number of mappings $f_{t}$ which are independent on $U F_{j}$ : clearly $s \leq p\left(F_{1}, \cdots, F_{n}\right)$. We may suppose that the
mappings $f_{k}$ are independent on $U F_{j}$ for $N-s+1 \leq k \leq N$, and then have, for each $t \leq N-s$, a relation (say)

$$
g_{t} \equiv f_{t}^{p_{t}} \prod_{k} f_{k}^{q_{k t}} \sim 1
$$

on $U F_{j}$, where $N-s+1 \leq k \leq N$. Thus $g_{t}$ is a mapping of $\cup A_{j}$ in $S^{1}$ which $\sim 1$ on each $A_{j}$; and, since clearly $p_{t} \neq 0$, the mappings $g_{t}(1 \leq t \leq N-s)$ are independent on $\cup A_{j}$.

Write $C_{0}=\mathrm{Cl}\left(S-\cup A_{j}\right)$; then $\operatorname{Fr}\left(C_{0}\right) \subset \cup F_{j}$, so that, from $1.3(7)$, each $g_{t}$ may be extended to a mapping (still denoted by $g_{t}$ ) of $S$ in $S^{1}$ such that $g_{t} \sim 1$ on $C_{0}$. Now define $C_{1}=A_{1}, C_{j}=\operatorname{Cl}\left[A_{j}-\left(A_{1} \cup A_{2} \cup \cdots \cup A_{j-1}\right)\right](2 \leq j \leq n)$; then the sets $C_{0}, C_{1}, \ldots, C_{n}$ are closed and cover $S$, and each $g_{t} \sim 1$ on each $C_{j}$. Let $Z=U\left(C_{j} \cap C_{k}\right)$, where $0 \leq j<k \leq n$; then $Z \subset U \operatorname{UFr}\left(A_{j}\right) \subset U F_{j}$, so that each $g_{t} \sim 1$ on $Z$. From l.3(9), the number $N-s$ of mappings $g_{t}$ is at most $r(S)$, and the theorem follows.
3.4. Remark. We remark that no inequality similar to Theorem 4, but in the reverse direction, can hold in general. For example, take $S$ to be the plane, and let $A$ be a circular disc and $B$ an inscribed convex polygon plus its interior; then $A, B$ are closed and connected, and $h(A, B)=0$, but $h[\operatorname{Fr}(A), \operatorname{Fr}(B)]$ can be arbitrarily large.

## 4. Frontiers of complementary components.

4.1. Definition. For any $A \subset S$, let $\left\{C_{\lambda}\right\}$ be the components of the complement of $A$, and write
(1) $c(A)=\sum b_{0}\left(\operatorname{Fr}\left(C_{\lambda}\right)\right)$,
with the usual convention that a vacuous sum is zero. [Thus $c(S)=0$, $c(0)=-1$.$] From [5, Th. 4] we have$
(2) $c(A)+b_{0}[\mathrm{Cl}(S-A)] \geq b_{0}[\operatorname{Fr}(A)]$,
and (a weaker statement unless $b_{0}[\mathrm{Cl}(S-A)]$ is infinite)
(3) $c(A) \geq b_{0}(\bar{A})$.

If $A$ is open, we evidently have equality in (2). (Note that (3) contains the well-known fact that, if $\bar{A}$ is not connected, at least one component of $\mathrm{Co}(A)$ has a disconnected frontier.)
4.2 Lemma. Let $C$ be a component of $S-A$, and let $U$ be an open set containing $\operatorname{Fr}(C)$. Then there exists an open set $V \supset \bar{A}$ such that $\bar{V} \cap \bar{C} \subset U$.

This follows from $[6, \S 6.1]$ applied to the sets $\bar{A} \bar{C}$; a direct proof is also easy.
4.3 Theorem 5. If $c(A) \geq n$, then there exists an open set $A^{*} \supset \bar{A}$ such that, for each set $B$ satisfying $A \subset B \subset A^{*}$, we have $c(B) \geq n$.

For if $c(A) \geq n$, then there exist finitely many components, say $C_{1}, C_{2}$, $\ldots, C_{m}$, of $\operatorname{Co}(A)$, such that $b_{0}\left[\operatorname{Fr}\left(C_{j}\right)\right] \geq n_{j}$ where $\sum n_{j} \geq n(1 \leq j \leq m)$. Thus, for each $j, \operatorname{Fr}\left(C_{j}\right)$ is a union of $n_{j}+1$ disjoint closed nonempty sets $F_{j k}\left(1 \leq k \leq n_{j}+1\right)$, and there exist open sets $U_{j k} \supset F_{j k}$ such that $\mathrm{Cl}\left(U_{j k}\right)$ $n \mathrm{Cl}\left(U_{j l}\right)=0(j \neq l)$. Let $U_{j}=\mathrm{U}_{k} U_{j k}$, an open set containing $\operatorname{Fr}\left(C_{j}\right)$; from the lemma in 4.2 , there exists an open set $V_{j} \supset \bar{A}$ such that $\mathrm{Cl}\left(V_{j}\right) \cap \mathrm{Cl}\left(C_{j}\right)$ $\subset U_{j}$. Take $A^{*}=\cap_{j} V_{j}$, and suppose that $B$ is any set satisfying $A \subset B \subset A^{*}$. Then, since $\cup_{k} F_{j k} \subset \bar{B} \cap \bar{C}_{j} \subset \cup_{k} U_{j k}$, we have $b_{0}\left(\bar{B} \cap \bar{C}_{j}\right) \geq n_{j}$. Now let $\left\{D_{j \mu}\right\}$ be those components of $\mathrm{Co}(B)$ which are contained in $C_{j}$, and write $E_{j}=U_{\mu} D_{j \mu}$. One readily verifies that $\operatorname{Fr}\left(E_{j}\right) \subset \bar{B} \cap \bar{C}_{j} \subset \mathrm{Cl}\left(S-E_{j}\right)$, and that $E_{j} \cup\left(\bar{B} \cap \bar{C}_{j}\right)=\bar{C}_{j}$; hence, from [5, Th. 4], $\sum_{\mu} b_{0}\left(\operatorname{Fr}\left(D_{j \mu}\right)\right) \geq b_{0}\left(\bar{B} \cap \bar{C}_{j}\right)$ $\geq n_{j}$, so that $c(B) \geq \sum_{j, \mu} b_{0}\left(\operatorname{Fr}\left(D_{j \mu}\right)\right) \geq \sum n_{j} \geq n$.

Corollary. We have $c(A) \leq c(\bar{A})$.
4.4. Extension of the Phragmén-Broumer theorem. This theorem, as extended in [5, Th. 5], can now be extended still further.

Theorem 6. For any set $A$, we have $c(A) \leq b_{0}(\bar{A})+r(S)$.
The proof is almost identical with that for the case in which $\bar{A}$ is connected, in $[5, \S 4.2]$; the difference arises from the fact that the sets $L, M$ there constructed need not here be connected. But we may assume without loss that $b_{0}(\bar{A})<\infty$, and have $b_{0}(L) \leq b_{0}(\bar{A})$ and $b_{0}(M) \leq b_{0}(\bar{A})$; hence, from $2.3(1)$, we have $b_{0}(L \cap M)<2 b_{0}(\bar{A})+r(S)$. Since $b_{0}(\bar{A})+1$ of the components of $L \cap M$ now arise from $\bar{A}$, the argument can be concluded in the same way as before.

Corollary 1. If $r(S)$ is finite, and $A$ is any subset of $S$ such that $\bar{A}$ has only a finite number of components, then all but at most a finite number of the components of $S-A$ have connected frontiers.

Corollary 2. If $S$ is unicoherent, then $c(A)=b_{0}(\bar{A})$; and, conversely, this equality is characteristic of unicoherence.
(This follows from 4.1 (2) and [5, Th. 5].)
4.5. Another extension. It has been shown in [5, Th. 5] that, conversely, Theorem 6 serves to characterize $r(S)$, even when restricted to the case in which $A$ is closed (or open) and connected. However, Theorem 6 can be restated in a slightly different though equally natural way, in which the converse question is more difficult.

Theorem 6a. For any set $A$, we have
(i) $b_{0}(\mathrm{Fr}(A)) \leq c(A)+b_{0}(\mathrm{Cl}(S-A)) \leq c(\bar{A})+b_{0}(\mathrm{Cl}(S-A))$

$$
\leq b_{0}(\operatorname{Fr}(A))+r(S)
$$

Conversely, if for some fixed (finite) $n$ we have
(ii) $c(A) \leq b_{0}(\operatorname{Fr}(A))+n$
whenever $A$ is nowhere dense, and if further
(iii) $S$ is metrizable, or $r(S)$ is finite, then $r(S) \leq n$.

The first inequality in (i) is a restatement of 4.1 (2), the second follows from Theorem 5, Corollary, and the third from Theorem 6 applied to $\bar{A}$, in view of the fact $[4, \delta 6.2]$ that $b_{0}(\bar{A})+b_{0}(\mathrm{Cl}(S-A)) \leq b_{0}(\operatorname{Fr}(A))$. For the converse, suppose that (ii) holds, but that $r(S)>n$. From [5, Th. 5a), there exists a closed connected set $A^{\prime}$ such that $S-A^{\prime}$ has only a finite number of (open) components $C_{1}, C_{2}, \ldots, C_{m}$, and $b_{0}\left(\operatorname{Fr}\left(A^{\prime}\right)\right)>m+n-1$; thus from [5, Th. 4], we have $\sum b_{0}\left(\operatorname{Fr}\left(C_{j}\right)\right)>n$. Suppose now that $r(S)$ is finite, and write $A=\operatorname{Fr}\left(A^{\prime}\right)$; thus $A$ is nowhere dense, and, from $2.3(2), b_{0}(A)<\infty$. Let $\left\{D_{\lambda}\right\}$ be the components of $\operatorname{Int}\left(A^{\circ}\right)$; then [5, Th. 4] we have $\sum b_{0}\left(\operatorname{Fr}\left(D_{\lambda}\right)\right] \geq b_{0}(A)$ $=b_{0}[\operatorname{Fr}(A)]$. But the components of $\operatorname{Co}(A)$ are precisely the sets $C_{j}, D_{\lambda}$; hence $c(A)>b_{0}[\operatorname{Fr}(A)]+n$, contradicting (ii).

If $r(S)=\infty$, the above argument still applies provided that $b_{0}\left[\operatorname{Fr}\left(A^{\prime}\right)\right]<\infty$. Hence we may assume $b_{0}\left[\operatorname{Fr}\left(A^{\prime}\right)\right]=\infty$, so that there must exist some $C_{j}$, say $C$, for which $b_{0}[\operatorname{Fr}(C)]=\infty$. Now, the complement (say) $F$ of $C$ is closed and connected. If it is assumed that $S$ is metrizable, then there exists a sequence of open sets $G_{n}$ such that $G_{n} \supset \mathrm{Cl}\left(G_{n+1}\right)(n=1,2, \ldots)$, and $\cap G_{n}=F$. Let $X=C-U \operatorname{Fr}\left(G_{n}\right)$; from a theorem of Hewitt [3], there exist disjoint sets $Y, Z$ such that $Y \cup Z=X$ and $\bar{Y}=\bar{Z}=\bar{X}=\bar{C}$. We take $A=C-Y$. Thus $\mathrm{Cl}(S-A)=S$; and $\operatorname{Fr}(A)=\bar{C}$, which is connected. But $\operatorname{Co}(A)$ can be separated, by one of the sets $\operatorname{Fr}\left(G_{n}\right)$, between $F$ and any given point of $Y$; thus one of the components of $\mathrm{Co}(A)$ is $F$ itself, and again (ii) is contradicted.

Corollary. If $S$ is unicoherent, and $\left\{C_{\lambda}\right\}$ are the components of an
arbitrary set $E$, then

$$
b_{0}\left[\operatorname{Fr}\left(C_{\lambda}\right)\right]+b_{0}(\bar{E})=b_{0}[\operatorname{Fr}(E)] ;
$$

and this property characterizes unicoherence among metrizable (locally connected and connected) spaces.

It would be interesting to know whether the extra hypotheses on $S$ imposed in (iii) are needed. It would be easy to replace them by others (for example, local compactness plus perfect normality).

## 5. Modified addition theorems.

5.1. A Modification. As in the case of two connected sets [5, Ths. 11 and 1la], special cases of Theorem 1 can be obtained under alternative hypotheses. As an example, we state:

Theorem 7. If $A$ and $B$ are any sets satisfying

$$
\operatorname{Fr}(A) \cap \operatorname{Fr}(B) \cap \operatorname{Fr}(A \cap B)=0,
$$

then

$$
b_{0}(A \cup B)+b_{0}(A \cap B) \leq b_{0}(A)+b_{0}(B)+r(S) ;
$$

and if there is finite equality here, then $A-B$ and $B-A$ are separated (so that Theorem 1 then in fact applies). ${ }^{9}$

The proof is a fairly straightforward generalization of that of [5, Th. 11], with 2.3 (l) replacing $[5, \S 7.4]$. The extension of Theorem 7 to $n$ sets, however, appears to present some difficulty.
5.2. Another Modification. A more interesting modification of Theorem 1 is the following, in which $r(S)$ does not enter explicitly; in some cases (in view of Theorem 6) it gives more information than does Theorem 1.

Theorem 8. If $A$ and $B$ are arbitrary sets such that

$$
\operatorname{Fr}(A) \cap \operatorname{Fr}(B) \cap \operatorname{Fr}(A \cup B)=0,
$$

then

$$
h(\bar{A}, \bar{B})+b_{0}(\bar{A}) \leq c(A) .
$$

[^7]Proof. Write $C=\mathrm{Cl}[\mathrm{Co}(A)]$, and apply [4, Th. 6b] to the closed sets $\bar{A} \cup \bar{B}$, $\bar{A} \cap \bar{B}, C$. We obtain
(1) $b_{0}(\bar{A} \cup \bar{B})+b_{0}(\bar{A} \cap \bar{B})+b_{0}(C) \leq b_{0}[\bar{B} \cup \operatorname{Fr}(A)]+b_{0}[\bar{B} \cap \operatorname{Fr}(A)]$.

From the frontier relation satisfied by the sets $A$ and $B$, it readily follows that $\operatorname{Fr}(A) \cap \operatorname{Co}(\bar{B})$ is closed, and thence that each component of $\operatorname{Fr}(A)$ which meets $\bar{B}$ is contained in $\bar{B}$. Hence we see that

$$
b_{0}[\bar{B} \cup \operatorname{Fr}(A)]+b_{0}[\bar{B} \cap \operatorname{Fr}(A)]=b_{0}(\bar{B})+b_{0}[\operatorname{Fr}(A)],
$$

and consequently

$$
\begin{equation*}
b_{0}(\bar{A} \cup \bar{B})+b_{0}(\bar{A} \cap \bar{B})+b_{0}(C) \leq b_{0}(\bar{B})+b_{0}[\operatorname{Fr}(A)] . \tag{2}
\end{equation*}
$$

But by $4.1(2)$, we have $b_{0}[\operatorname{Fr}(A)] \leq b_{0}(C)+c(A)$. Thus, provided that $b_{0}(C)$ is finite, we have proved
(3) $b_{0}(\bar{A} \cup \bar{B})+b_{0}(\bar{A} \cap \bar{B}) \leq b_{0}(\bar{B})+c(A)$,
from which the theorem follows immediately.
To complete the proof, we deduce that (3) continues to hold even when $b_{0}(C)=\infty$; and in doing so, we may assume that $b_{0}(\bar{B})+c(A)<\infty$. Define $B^{*}$ to be the union of those components of $\bar{B}$ which meet $\bar{A}$, and $A^{*}$ to be the union of $A$ with all components of $\operatorname{Co}(A)$ which have connected frontiers. It is easy to verify that

$$
\operatorname{Fr}\left(A^{*}\right) \cap \operatorname{Fr}\left(B^{*}\right) \cap \operatorname{Fr}\left(A^{*} \cap B^{*}\right)=0,
$$

and that, since $c(A)<\infty, \mathrm{b}_{0}\left[\mathrm{Co}\left(A^{*}\right)\right]$ is finite. Hence (3) holds for the sets $A^{*}, B^{*}$; and it is a routine matter to deduce that (3) also holds for $A$ and $B^{*}$, and thence finally for $A$ and $B$.

There is no difficulty in extending this theorem to any number of sets; for example, (2) can be extended to the following property, valid in an arbitrary topological space $S$ (and generalizing [4, §7.4(1)]):
(4) If $A_{1}, \cdots, A_{m}, B_{1}, \cdots, B_{n}$ are arbitrary sets such that

$$
\operatorname{Fr}\left(A_{j}\right) \cap \operatorname{Fr}\left(B_{k}\right) \cap \operatorname{Fr}\left(A_{j} \cup B_{k}\right)=0 \quad(1 \leq j \leq m, 1 \leq k \leq n),
$$

and $C_{j}=\mathrm{Cl}\left[\mathrm{Co}\left(A_{j}\right)\right]$, then

$$
\begin{aligned}
& \sum b_{0}\left[X _ { h } \left(\overline{A_{1}}, \cdots,\right.\right. \bar{A}_{m}, \\
& \bar{B}_{1}\left.\left., \cdots, \bar{B}_{n}\right)\right]+\sum b_{0}\left[X_{j}\left(C_{1}, \ldots, C_{m}\right)\right] \\
& \leq \sum b_{0}\left[X_{j}\left[\operatorname{Fr}\left(A_{1}\right), \ldots, \operatorname{Fr}\left(A_{m}\right)\right]\right\} \\
&+\sum b_{0}\left[X_{k}\left(\bar{B}_{1}, \cdots, \bar{B}_{n}\right)\right]+m b_{0}(S)
\end{aligned}
$$

the ranges of summation being $1 \leq h \leq m+n, 1 \leq j \leq m, 1 \leq k \leq n$; and (3) can be extended similarly.
5.3. An Inclusive Result. The next theorem includes both Theorem 1 and the extended Phragmén-Brouwer theorem (Theorem 6) as special cases. We shall need the following lemma.

Lemma. If $G$ is a set with only finitely many components, then there exists a finite set of points $x_{1}, x_{2}, \cdots, x_{q} \in \operatorname{Fr}(G)$ such that

$$
b_{0}\left[G \cup\left(x_{1}\right) \cup \cdots \cup\left(x_{q}\right)\right]=b_{0}(\bar{G})
$$

For if $G$ has components $G_{1}, G_{2}, \ldots, G_{s}$, we have only to take at least one point $x_{j}$ in every nonempty set $\bar{G}_{\lambda} \cap \bar{G}_{\mu} \quad(\lambda \neq \mu)$.
5.4. Theorem 9. Let $A_{1}, A_{2}, \cdots, A_{n}$ be any subsets of $S$ having separated differences and such that $A_{j} \cap A_{k}$ and $\mathrm{Co}\left(A_{j} \cup A_{k}\right)$ are separated whenever $j \neq k$; and suppose that no point belongs to $A_{j}$ for more than $m$ values of $j$, where $2 \leq m \leq n$. Then
(1) $h\left(A_{1}, \ldots, A_{n}\right)+c\left(\bar{X}_{1}\right)+c\left(\bar{X}_{2}\right)+\cdots+c\left(\bar{X}_{m-1}\right)$

$$
\leq b_{0}\left(\bar{X}_{1}\right)+b_{0}\left(\bar{X}_{2}\right)+\cdots+b_{0}\left(\bar{X}_{m-1}\right)+(m-1) r(S),
$$

where $X_{j}=X_{j}\left(A_{1}, \cdots, A_{n}\right)$. Further, if there is finite equality in (1), then, for each $q \leq n-1$, for each set $J$ of $q+1$ distinct suffixes $j_{1}$, $j_{2}, \cdots, j_{q+1}$ between 1 and $n$, and for each component $E$ of $X_{q}$, we have
(2) $\cap\left\{\operatorname{Fr}\left(A_{j} \cap E\right) \mid j \in J\right\} \subset E$.

To prove (1), we may assume throughout that $r(S)$ and $\sum b_{0}\left(A_{j}\right)$ are finite; it then follows from Theorems 1 and 6 that the numbers $b_{0}\left(X_{j}\right)$, $b_{0}\left(\overline{X_{j}}\right)$, and $c\left(X_{j}\right)$ are also finite. Further, we may obviously suppose that $X_{m-1} \neq 0$ (otherwise (1) would be derived with a smaller value of $m$ ). Again, by using the method of approximation, we may assume in addition that the sets $A_{j}$ are all open and, by $1.3(2)$, satisfy
(3) $\operatorname{Fr}\left(A_{j}\right) \cap \operatorname{Fr}\left(A_{k}\right) \cap \operatorname{Fr}\left(A_{j} \cap A_{k}\right)=0$

For, in the general case, we apply 1.3 (3) to replace the sets $A_{j}$ by slightly larger relatively connected sets $A_{j}{ }^{*}$ having the same incidences and satisfying (3); and, in view of Theorem 5, the truth of (1) for the sets $A_{j}^{*}$ will imply (1) for the sets $A_{j}$.

From (3) and $1.3(1)$, the open sets $X_{j}$ satisfy
(4) $X_{1} \supset \bar{X}_{2} \supset X_{2} \supset \overline{X_{3}} \supset \cdots \supset X_{m+1}=0$.

We shall define inductively, for $j=1,2, \cdots, m-1$, open sets $G_{j}$ consisting of a finite number of components $C_{j k}$ of $\operatorname{Co}\left(\bar{X}_{j}\right)$, and open sets $V_{j} \supset \operatorname{Fr}\left(G_{j}\right)$, such that ${ }^{11}$
(5) $G_{j} \cup V_{j} \subset G_{k}$ whenever $j<k$;
$\bar{V}_{j} \subset X_{j-1} ; \bar{V}_{j} \cap \bar{X}_{j+1}=0 ; \bar{V}_{j} \cap \bar{V}_{k}=0$ if $j \neq k ;$
$\operatorname{Fr}\left(V_{j}\right) \cap \operatorname{Fr}\left(A_{k}\right)=0$ (for all $\left.j, k\right) ;$ and $\operatorname{Fr}\left(V_{j}\right) \cap \operatorname{Fr}\left(X_{j}\right)=0$.
Further,
(6)

$$
\begin{aligned}
b_{0}\left(V_{j}\right)<\infty, b_{0}\left(X_{j} \cup V_{j}\right) & \leq b_{0}\left(\bar{X}_{j}\right), \text { and } \\
b_{0}\left(V_{j} \cap G_{j}\right) & \geq c\left(\overline{X_{j}}\right)+b_{0}\left(G_{j}\right)
\end{aligned}
$$

For suppose this done for all $j<p$, where $1<p<m$. Define $G_{p}$ to be the union of all those components of $\operatorname{Co}\left(\bar{X}_{p}\right)$ which either (a) have disconnected frontiers, or (b) meet $G_{p-1} \cup V_{p-1}$. Since $G_{p-1} \cup V_{p-1}$ $\subset \operatorname{Co}\left(\bar{X}_{p}\right)$, this gives $G_{p-1} \cup V_{p-1} \subset G_{p}$; and since further

$$
b_{0}\left(G_{p-1} \cup V_{p-1}\right)<\infty,
$$

Theorem 6, Corollary 1, shows that $b_{0}\left(G_{p}\right)<\infty$. Let $G_{p}$ consist of the components $C_{p k}$ of $\operatorname{Co}\left(\bar{X}_{p}\right)\left(k=1,2, \cdots, n_{p}\right)$; thus

$$
\sum_{k} b_{0}\left[\operatorname{Fr}\left(C_{p k}\right)\right]=c\left(\bar{X}_{p}\right)
$$

Hence, if the components of $\operatorname{Fr}\left(C_{p k}\right)$ are denoted by $F_{p k l} \quad(l=1,2$, $\left.\ldots, n_{p k}\right)$, we have $\sum_{k}\left(n_{p k}-1\right)=c\left(\bar{X}_{p}\right)$. For fixed $p$ and $k$, there exist open sets $N_{p k l} \supset F_{p k l}$ with disjoint closures (for varying $l$ ); and it follows from the lemma in 4.2 that an open set $U_{p}$ exists such that $U_{p} \supset \operatorname{Fr}\left(X_{p}\right)$ and $U_{p} \cap \mathrm{Cl}\left(C_{p k}\right) \subset \mathrm{U}_{l} N_{p k l}$. We may further suppose, from (4), that $\bar{U}_{p} \subset X_{p-1}, \bar{U}_{p} \cap \bar{X}_{p+1}=0$, and $\bar{U}_{p} \cap \bar{V}_{p-1}=0$. It readily follows that $\operatorname{Fr}\left(U_{p}\right) \cap \operatorname{Fr}\left(A_{k}\right)=0$ for each $k(1 \leq k \leq n)$. Again, from
the lemma in 5.3 , there exists a finite set $Q_{p} \subset \operatorname{Fr}\left(X_{p}\right)$ such that

$$
b_{0}\left(X_{p} \cup Q_{p}\right)=b_{0}\left(\bar{X}_{p}\right)
$$

Let $V_{p}=$ union of those components of $U_{p}$ which meet $Q_{p} \cup \cup F_{p k l}$; clearly $b_{0}\left(V_{p}\right)$ is finite; and, since $\operatorname{Fr}\left(V_{p}\right) \subset \operatorname{Fr}\left(U_{p}\right)$, it is easily seen that (5) continues to hold when $j=p$. Also the sets $V_{p} \cap C_{p k} \cap N_{p k l}$ are (for varying $k$ and $l$ ) all pairwise separated and nonempty; hence the number of components of $V_{p} \cap G_{p}$ is at least $\sum_{k} n_{p k}=c\left(\bar{X}_{p}\right)+n_{p}$, so that (6) holds.

To start the induction, we take $G_{1}$ to consist of the components of $\mathrm{Co}\left(\bar{X}_{1}\right)$ with disconnected frontiers; the rest of the construction is exactly as in the general case. Thus (5) and (6) hold for $j=1,2, \cdots, m-1$. We remark that it follows trivially from (5) that
(7) $\operatorname{Fr}\left(V_{j}\right) \cap \operatorname{Fr}\left(G_{k}\right)=0$ whenever $j \neq k$, and $V_{j} \cap G_{k}=0$ if $j>k$.

Now consider the "elementary symmetric sets"

$$
Y_{j}=X_{j}\left(G_{1}, G_{2}, \cdots, G_{m-1}, V_{1}, V_{2}, \cdots, V_{m-1}\right) \quad(1 \leq j \leq 2 m-2)
$$

and

$$
\begin{aligned}
Z_{k}=X_{k}\left(A_{1}, A_{2}, \cdots, A_{n}, G_{1}, \cdots, G_{m-1}, V_{1}, \cdots,\right. & \left.V_{m-1}\right) \\
& (1 \leq k \leq 2 m+n-2) .
\end{aligned}
$$

Using (5) and (7), we obtain
(8) $Y_{1}=G_{m-1} \cup V_{m-1}, Y_{j}=G_{m-j} \cup V_{m-j} \cup\left(G_{m-j+1} \cap V_{m-j+1}\right)$ if $2 \leq j \leq m-1 ; Y_{m}=G_{1} \cap V_{1} ;$ and $Y_{j}=0$ if $j>m$.

Thus, since $Z_{k}=X_{k} \cup \cup\left(X_{k-p} \cap Y_{p}\right) \cup Y_{k}$, we find:
(9) $Z_{k} \neq 0$ if $1 \leq k \leq m ; Z_{m}=X_{m} \cup \cup\left(X_{p} \cap V_{p}\right) \cup \cup\left(G_{q} \cap V_{q}\right)$ $(p, q=1,2, \cdots, m-1)$; and $Z_{k}=0$ if $k>m$.

Now the open sets $A_{1}, A_{2}, \cdots, A_{n} ; G_{1}, \cdots, G_{m-1} ; V_{1}, \cdots, V_{m-1}$ satisfy the hypotheses of Theorem 1 , since this is true of $A_{1}, \cdots, A_{n}$ from (3), while we readily verify that

$$
\operatorname{Fr}\left(A_{j}\right) \cap \operatorname{Fr}\left(G_{k}\right) \cap \operatorname{Fr}\left(A_{j} \cap G_{k}\right)=0
$$

and that in all the remaining cases the sets have disjoint frontiers. Hence

$$
\begin{align*}
& \sum b_{0}\left(Z_{p}\right) \leq \sum b_{0}\left(A_{j}\right)+\sum b_{0}\left(G_{k}\right)+\sum b_{0}\left(V_{k}\right)+(m-1) r(S)  \tag{10}\\
& (1 \leq p \leq n+2 m-2,1 \leq j \leq n, 1 \leq k \leq m-1) .
\end{align*}
$$

But (9) shows that

$$
\sum b_{0}\left(Z_{p}\right) \geq b_{0}\left(X_{m}\right)+\sum b_{0}\left(X_{k} \cap V_{k}\right)+\sum b_{0}\left(V_{k} \cap G_{k}\right)+m-n .
$$

Also

$$
\begin{aligned}
b_{0}\left(X_{k} \cap V_{k}\right) & \geq b_{0}\left(X_{k}\right)+b_{0}\left(V_{k}\right)-b_{0}\left(X_{k} \cap V_{k}\right) \quad(\text { cf. [4, §6.2] }) \\
& \geq b_{0}\left(X_{k}\right)+b_{0}\left(V_{k}\right)-b_{0}\left(\bar{X}_{k}\right),
\end{aligned}
$$

and, from (6).

$$
b_{0}\left(V_{k} \cap G_{k}\right) \geq c\left(\bar{X}_{k}\right)+b_{0}(G) .
$$

Thus finally, since all the numbers involved here are finite, (1) follows from (10).
5.5. The Case of Finite Equality. Suppose now that there is finite equality in (1) above, and that a point $y$ exists in (say)

$$
\operatorname{Fr}\left(A_{1} \cap E\right) \cap \operatorname{Fr}\left(A_{2} \cap E\right) \cap \ldots \cap \operatorname{Fr}\left(A_{p+1} \cap E\right)-E,
$$

where $E$ is a component of $X_{p}$; thus $y \notin X_{p}$. It is easy to see that we may assume without loss of generality that $p \leq m-1$ and that the sets $A_{j}$ are all open. Clearly $y \in \operatorname{Fr}\left(X_{p}\right)$; thus we may carry out the preceding construction in such a way that $y \in Q_{p} \subset V_{p}$. But, from the way in which (1) was derived from (10), we must now have $h\left(X_{p}, V_{p}\right)=0$, so that the component $W$ of $V_{p}$ which contains $y$ must meet $E$ in a connected set; consequently, since $W \cap \mathrm{Cl}\left(X_{p+1}\right)$ $=0$, it follows that $W \cap E$ meets one and only one of the sets

$$
A_{J}\left(=\cap\left\{A_{j}, j \in J\right\}\right)
$$

with $|J|=p$. Since $\mathscr{W}$ meets $A_{1} \cap E$, we have $1 \in J$; similarly $2 \in J, \cdots$, and $(p+1) \in J$, giving a contradiction.
5.6. Remarks. We observe that the preceding results contain those concerning modified addition theorems in [5, Ths. 11 and 1la]. For, in the first place, l.3(1) together with an "approximation" argument shows that the relation
(2) above is equivalent to the apparently stronger relation

$$
\begin{equation*}
\cap\left\{\operatorname{Fr}\left(A_{j} \cap E\right)\right\} \subseteq \operatorname{Cl}\left(X_{p+1}\right) \tag{2a}
\end{equation*}
$$

$$
(j \in J,|J|=p+1, p<n)
$$

(In fact, the left side here is contained in

$$
\left.\operatorname{Fr}\left(X_{p+1}\right) \cup \operatorname{Fr}\left(X_{p+2}\right) \cup \cdots \cup \operatorname{Fr}\left(X_{m}\right) .\right)
$$

Hence if $A_{1}, \cdots, A_{n}$ also satisfy the condition (slightly stronger than in Theorem 9) that $\operatorname{Fr}\left(A_{j} \cap A_{k}\right) \cap \operatorname{Fr}\left(A_{j} \cup A_{k}\right)=0(j \neq k)$, finite equality in Theorem 9 will imply, again from $1.3(1)$, that

$$
\begin{equation*}
\cap\left\{\operatorname{Fr}\left(A_{j} \cap E\right)\right\} \subset \operatorname{Int}(E) \tag{2b}
\end{equation*}
$$

$$
(j \in J,|J|=p+1, p<n)
$$

a relation which is slightly stronger than (2). And if the sets $A_{j}$ satisfy the even stronger condition

$$
\operatorname{Fr}\left(A_{j}\right) \cap \operatorname{Fr}\left(A_{k}\right) \cap \operatorname{Fr}\left(A_{j} \cap A_{k}\right)=0 \quad(j \neq k),
$$

it can be deduced from (2b) that

$$
\begin{equation*}
\cap\left\{\operatorname{Fr}\left(A_{j} \cap E\right)\right\}=0 \quad \text { if } 2 p \geq n \quad(j \in J,|J|=p+1, p<n) \tag{2c}
\end{equation*}
$$

Finally, if there is finite equality in Theorem l, then there will be finite equality in Theorem 9 , for $c\left(\bar{X}_{j}\right) \geq b_{0}\left(\bar{X}_{j}\right)$, by 4.1 (3); and thus the above considerations will apply.
5.7. Other Inequalities. Many other inequalities can be derived from Theorem 9; for example :

Theorem 9a. Under the hypotheses of Theorem 9, we have
(i) $c\left(\bar{X}_{1}\right)+c\left(\bar{X}_{2}\right)+\cdots+c\left(\bar{X}_{m-1}\right)+b_{0}\left(X_{m}\right)+b_{0}\left(X_{m+1}\right)+\cdots+b_{0}\left(X_{n}\right)$

$$
\leq \sum b_{0}\left(A_{j}\right)+(m-1) r(S)
$$

Further, if there is finite equality in (i), we have

$$
\text { (ii) } X_{p}\left(\overline{A_{1}}, \overline{A_{2}}, \cdots, \overline{A_{n}}\right)=\bar{X}_{p} \quad(1 \leq p \leq m) .^{12}
$$

Proof. Relation (i) is a trivial consequence of Theorem 9, (1), since

$$
b_{0}\left(\bar{X}_{j}\right) \leq b_{0}\left(X_{j}\right)
$$

12 Condition (ii) need not hold for $p>m$.

Suppose there is finite equality in (i); as before it will suffice, by an approximation argument, to prove (ii) assuming that the sets $A_{j}$ are open. Now, finite equality in (i) implies that Theorem 9 , (2), holds, and also that $b_{0}\left(\overline{X_{j}}\right)=b_{0}\left(X_{j}\right)$ for all $j<m$. If $J, K$ denote sets of $p$ distinct suffixes between $l$ and $n$, and $J \neq K$, we find (writing $A_{J}=\cap\left\{A_{j} \mid j \in J\right\}$ ) that

$$
\bar{A}_{J} \cap \bar{A}_{K} \subset \bar{X}_{p+1} \quad \text { if } p<m
$$

This includes (ii) when $p=2$; and (ii) follows in general by an easy induction over $p$.
5.8. Geometrical Considerations. To illustrate the geometrical content of these theorems, we consider the case of two sets in more detail.

Theorem 10. Let $A$ and $B$ be sets, neither of which contains the other, having separated differences and connected complements, and suppose that $A \cap B$ and $\operatorname{Co}(A \cup B)$ are separated. Then

$$
b_{0}(A \cap B)+b_{0}(\operatorname{Co}(A \cup B)) \leq b_{0}(A)+b_{0}(B)+r(S)-1 .
$$

If there is finite equality here, and further $\operatorname{Fr}(A) \cap \operatorname{Fr}(B) \cap \operatorname{Fr}(A \cup B)=0$, then each component of $\mathrm{Co}(A \cup B)$ has a frontier consisting of exactly two components.

Proof. We can assume that $r(S)$ is finite. Write $P=\operatorname{Co}(A), Q=\operatorname{Co}(B)$; then $P$ and $Q$ are connected, so that, from Theorem $1, b_{0}(P \cap Q)$ is finite. Let $P \cap Q(=\operatorname{Co}(A \cup B))$ have components $H_{1}, H_{2}, \cdots, H_{n}$. Then
(1) $A$ and $H_{j}$ are not separated,
since otherwise $Q=\left\{(Q \cap A) \cup\left(P \cap Q-H_{j}\right)\right\} \cup H_{j}$, a union of two nonempty separated sets. Similarly $B$ and $H_{j}$ are not separated. Hence
(2) $\operatorname{Fr}\left(H_{j}\right)$ meets both $\operatorname{Fr}(A)$ and $\operatorname{Fr}(B)$.

Let $A^{*}=A \cup(P \cap Q), B^{*}=B \cup(P \cap Q)$; from (1), $A^{*}$ is connected relative to $A$, so that $b_{0}\left(A^{*}\right) \leq b_{0}(A)$, and similarly $b_{0}\left(B^{*}\right) \leq b_{0}(B)$. It is easy to see that $A^{*}-B^{*}$ and $B^{*}-A^{*}$ are separated, and that $A^{*} \cap B^{*}$ and $\operatorname{Co}\left(A^{*} \cup B^{*}\right)$ are separated; hence 2.3 (1) gives

$$
b_{0}\left(A^{*} \cap B^{*}\right) \leq b_{0}(A)+b_{0}(B)+r(S) .
$$

But $A^{*} \cap B^{*}=(A \cap B) \cup C o(A \cup B)$, a union of two separated sets; thus the first part of the theorem follows.

From Theorem 9a and Theorem 5, Corollary, we also have

$$
b_{0}(A \cap B)+\sum b_{0}\left(\operatorname{Fr}\left(H_{j}\right)\right) \leq b_{0}(A)+b_{0}(B)+r(S)
$$

If further $\operatorname{Fr}(A) \cap \operatorname{Fr}(B) \cap \operatorname{Fr}(A \cup B)=0$, (2) shows that $b_{0}\left[\operatorname{Fr}\left(H_{j}\right)\right]$ $\geq 1$ for each $j$. Hence finite equality in Theorem 10 requires $b_{0}\left[\operatorname{Fr}\left(H_{j}\right)\right]$ $=1$ for each $j$, and the proof is complete.

Corollary. If $A$ and $B$ are simple sets ${ }^{\mathbf{1 3}}$ with disjoint frontiers, and neither $A$ nor $B$ contains the other, then $b_{0}(A \cap B)+b_{0}[\operatorname{Co}(A \cup B)] \leq r(S)-1$; and if there is finite equality here, then each component of $A \cap B$ or of $\operatorname{Co}(A \cup B)$ has a frontier with exactly two components.

This follows on applying Theorem 10 first to $A, B$ and then to $\operatorname{Co}(A), \operatorname{Co}(B)$. If $S$ is unicoherent, the first part of this corollary reduces to [4, §4.5].

## 6. Simple sets with disjoint frontiers.

6.1. Finitely Multicoherent Spaces. Throughout this section, we shall assume that $r(S)$ is finite.

Theorem 11. Let $A_{1}, A_{2}, \cdots, A_{n}$ be simple ${ }^{\mathbf{1 4}}$ subsets of $S$, every two of which meet, and which have disjoint frontiers. Then there exist $N$ or fewer of the sets $A_{j}$ whose union is $\cup_{1}^{n} A_{j}$, where

$$
\begin{aligned}
& N=2 r(S) \text { if } r(S)>1, \text { or if } r(S)=1 \text { and } \cap A_{j} \neq 0 . \\
& N=3 \text { if } r(S)=1 \text { and } \cap A_{j}=0, \text { and } \\
& N=2 \text { if } r(S)=0
\end{aligned}
$$

These values of $N$ are the smallest possible.
It is easy to see by examples (it suffices to take $S$ to be a linear graph) that no smaller values of $N$ are possible in general. To prove the rest of the theorem, we need two graph-theoretic lemmas.
6.2. Lemma 1. Let $G$ be a connected linear graph having no end-points, and let $E_{1}, E_{2}, \ldots, E_{n}$ be closed connected subsets of $G$, every two of which meet. If $r(G)>1$, or if $r(G)=1$ and $\cap E_{j} \neq 0$, then $\cup E_{j}$ is the union of $2 r(G)$ or fewer of the sets $E_{j}$; if $r(G)=1$ and $\cap E_{j}=0$, then $\cup E_{j}$ is the union of at most 3 sets $E_{j}$.

[^8]The proof is by induction over $r(G)$. If $r(G) \leq 2$, the lemma can be verified by inspection of the possible graphs $G$. Suppose, then, that $r(G) \geq 3$, and that the lemma is true for all graphs of smaller degree of multicoherence but not for $G$, and that $n$ is the smállest number of sets for which the lemma fails for $G$. Thus no $E_{j}$ is contained in the union of the others.

From $G$ we derive a homeomorphic graph $G^{*}$ by suppressing all vertices of order 2; the (open) 1-cells of $G^{*}$ will thus be the components of $G$ minus its vertices of orders other than 2 ; we call them the "maximal 1-cells" of $G$. (A maximal l-cell may have coincident end-points.) We consider three cases:
(1) If $G$ has a cut-point $R$ which is not a vertex of $G$, let $P Q$ be the maximal l-cell of $G$ which contains $R$; thus here $P \neq Q$, and $G-P Q$ is a union of two disjoint, closed connected nonempty subgraphs $H, K$, neither of which has an end-point. From [6, §3.2(1)], we see that $r(H)+r(K)=r(G)$, while, since $G$ has no end-points, $r(H) \geq 1$ and $r(K) \geq 1$. For the moment we assume that neither $r(H)$ nor $r(K)$ is 1 . Write $E_{j}{ }^{\prime}=E_{j} \cap H, E_{j}{ }^{\prime \prime}=E_{j} \cap K$; it is easy to see that these sets are closed and connected, though possibly empty. Further, every two nonempty sets $E_{j}$ ' must meet, since both must contain $P$ unless one of the corresponding sets $E_{j}$ is contained in $H$. Hence the hypothesis of induction applies to $H$ and the nonempty sets $E_{j}^{\prime}$, and $\cup E_{j}^{\prime}$ must be contained in the union of at most $2 r(H)$ sets $E_{j}$. Similarly $U E_{j}$ " is contained in the union of at most $2 r(K)$ sets $E_{j}$. Thus we obtain at most $2 r(G)$ sets $E_{j}$ in all, which together contain $\cup E_{j}{ }^{\prime} \cup U E_{j}{ }^{\prime \prime}$; further, their union is connected and so contains $P Q$ and thus $U E_{j}$, unless $U E_{j}^{\prime}$ or $U E_{j}^{\prime \prime}$ is empty.

If $U E_{j}{ }^{\prime \prime}$, say, is empty but $U E_{j}{ }^{\prime} \neq 0$, it is easy to see that at most $2 r(H)+1$ $<2 r(G)$ sets $E_{j}$ will suffice, namely those selected to contain $U E_{j}^{\prime}$, together with the set $E_{j}$ which contains the largest subarc of $P Q$. If $U E_{j}^{\prime}=U E_{j}{ }^{\prime \prime}=0$, all the sets $E_{j}$ are contained in $P Q$, and two of them will suffice.

If $r(H)$, say, is 1 (so that $H$ is a circle), the above argument needs modification only if one of the given sets is contained in $H-(P)$; we leave the details to the reader.
(2) If $G$ has a cut-vertex $R$, but no cut-point other than a vertex, the argument is essentially the same as before, with $P Q$ degenerating to $R$.
(3) Finally, if $G$ has no cut-points, pick $x \in E_{1}-\left(E_{2} \cup \cdots u E_{n}\right)$; replacing $x$ by a sufficiently nearby point if necessary, we can suppose that $x$ is not a vertex and so belongs to a unique maximal l-cell $P Q$ of $G$. Here $P \neq Q$, since $G$ has no cut-points, and the subgraph $H=G-P Q$ is connected and has no endlines. We easily find $r(H)=r(G)-1$. Write $E_{j}^{\prime}=E_{j} \cap H$; as before, at most $2 r(H)$ of the sets $E_{2}, \cdots, E_{n}$, say $E_{2}, \cdots, E_{m}(m \leq 2 r(H)+1)$, must contain

UE $j^{\prime}(j \geq 2)$. The connected set $E_{1} \cup E_{2}$ joins $x$ to $H$ (for we may clearly assume $U E_{j}^{\prime} \neq 0$ ), and so contains one of the $\operatorname{arcs} P x, Q x$, say $P x$. If none of $E_{m+1}, \cdots, E_{n}$ meets $Q x$, the $m$ sets $E_{1}, E_{2}, \cdots, E_{m}$ contain $\cup E_{j}$. If $Q x$ n $\left(E_{m+1} \cup \ldots \cup E_{n}\right) \neq 0$, let $y$ be its point on $Q x$ closest to $x$, and let $y \in E_{k}$; then the connected set $E_{2} \cup E_{k}$ joins $y$ to $H$ without containing $x$, and so contains $Q X$; thus the $m+1$ sets $E_{1}, E_{2}, \cdots, E_{m}, E_{k}$ contain $U E_{j}$. Since $m+1$ $\leq 2 r(G)$, the proof is complete.
6.3. Lemma 2. Let $B_{1}, B_{2}, \cdots, B_{n}$ be $n$ simple closed subsets of a connected linear graph $G$, every two of which meet. If $r(G)>1$, or if $r(G)=1$ and $\cap B_{j} \neq 0$, then $\cup B_{j}$ is the union of $2 r(G)$ or fewer of the sets $B_{j}$; if $r(G)=1$ and $\cap B_{j}=0$, then $\cup B_{j}$ is the union of at most 3 sets $B_{j}$.

As before, we may assume that the lemma is false, and that $n$ is the smallest number of sets for which it fails; thus no $B_{j}$ is contained in the union of the others. Define a "maximal end-line" $P Q$ of $G$ to be a maximal l-cell $P Q$ of $G$ in which $Q$ is an end-point of $G$; thus $P \neq Q$. If $B_{1}$, say, meets a maximal endline $P Q$ which it does not contain, then (being closed and simple) $B_{1}$ must be either a closed arc $x Q$, where $x \in P Q$, or the closure of the complement in $G$ of such an arc. In the latter case, it is clear that $B_{1}$ together with one other set $B_{j}$ will contain the rest; in the former case, we see similarly that either $B_{1} \cup B_{2}$ $=G$, or $B_{2} \supset B_{1}$, or $B_{1} \supset B_{2}$-all of which are excluded. This proves, then, that each $B_{j}$ contains all maximal end-lines of $G$ which it meets. Let $H$ be the graph obtained from $G$ by removing all end-points and maximal end-lines, and write $E_{i}=B_{i} \cap H$. On applying Lemma 1 to the sets $E_{1}, \cdots, E_{n}$ in the graph $H$, we see that $U E_{j}$ is the union of the desired number of sets $E_{j}$; the analogous conclusion for the sets $B_{j}$ follows.
6.4. Proof of Theorem 11. We shall consider only the case $r(S)>1$ explicitly; the modifications needed when $r(S)=1$ will be obvious, and the case $r(S)=0$ is covered by [4, §4.5]. It will thus suffice to prove that, if $n>2 r(S)$ $\geq 4$, one of the sets $A_{j}$ is contained in the union of the others. Consider the $2^{n}$ intersections $Y_{k}=\bar{D}_{1} \cap \bar{D}_{2} \cap \ldots \cap \bar{D}_{n}\left(1 \leq k \leq 2^{n}\right)$, where each $D_{j}$ takes the two values $A_{j}, \mathrm{Co}\left(A_{j}\right)$, in all possible combinations. The sets $Y_{k}$ are closed and cover $S$; and, since the sets $\operatorname{Fr}\left(A_{j}\right)$ are disjoint, no three of them have a common point. Further, from Theorem $1, b_{0}\left(Y_{k}\right)$ is finite, and so the sets $Y_{k}$ are of finite incidence. Let $G$ denote the modified nerve (cf. [6]) of the sets $Y_{k}$; as in [ $6, \S 6.4$ ], $G$ is connected and $r(G) \leq r(S)$. Let $B_{p}$ denote the subgraph of $G$ consisting of (i) all vertices which correspond to intersections $Y_{k}$ in which the pth "factor" $D_{p}$ is $A_{p}$, and (ii) all edges of $G$ both of whose end-points
have been assigned to $B_{p}$. Let $C_{p}$ be defined similarly, but with $\operatorname{Co}\left(A_{p}\right)$ replacing $A_{p}$. Thus, for each $p(1 \leq p \leq n), B_{p}$ and $C_{p}$ are disjoint subgraphs which together contain all the vertices of $G$; and it is easy to see that $B_{p}$ and $C_{p}$ are connected, since $A_{p}$ and $\mathrm{Cl}\left[\mathrm{Co}\left(A_{p}\right)\right]$ are. Hence $B_{1}, B_{2}, \ldots, B_{n}$ are simple closed subsets of $G$. Further, if $p \neq q, B_{p}$ and $B_{q}$ have at least a common vertex. Thus, by Lemma 2, one of the sets $B_{p}$ is contained in the union of the others; say $B_{1} \subset B_{2} \cup \ldots \cup B_{n}$. It readily follows that $A_{1} \subset A_{2} \cup \cdots \cup A_{n}$, whence the proof is completed.
6.5. Corollary. For any collection of more than $2 r(S)$ simple subsets of $S$ with disjoint frontiers, the union of some two of the sets contains the intersection of the rest.
6.6. Further Results. Evidently the method which was employed to prove Theorem 11 is of more general applicability; it shows, roughly speaking, that the incidences of a system of sets with disjoint frontiers are no worse than if $S$ were a linear graph of the same degree of multicoherence. In the same way we may prove:

Theorem lla. Let $A_{1}, A_{2}, \cdots, A_{n}$ be $n$ simple subsets of $S$, every two of which meet, and which have disjoint frontiers. If $n$ is large enough compared with $r(S)$ (assumed finite), then some $A_{j}$ is contained in the union of two others.
(Note that no $A_{j}$ need be contained in one other, irrespective of how large $n$ is.) Here the determination of the "best" bound for $n$ seems to be difficult: it can be shown, however, that, disregarding the trivial case $r(S)=0$, it lies between $\exp \left\{\exp \left[c_{1} r(S)\right]\right\}$ and $\exp \left\{\exp \left[c_{2} r(S)\right]\right\}$, where $c_{1}, c_{2}$ are positive constants.

Another related theorem, proved in a similar way, is:
Theorem llb. Let $A_{1}, A_{2}, \cdots, A_{n}$ be connected subsets of $S$ such that $b_{0}\left[\operatorname{Co}\left(A_{j}\right)\right] \leq q(j=1,2, \cdots, n)$. Suppose that every two of the sets $A_{j}$ meet, and that they have disjoint frontiers. Then there exists a function $N$ of $q$ and $r(S)$ (independent of $n$ ) such that $\cup A_{j}$ is contained in the union of $N$ or fewer of the sets $A_{j}$.

It is easy to show by examples that, with $q \geq 1$, we have

$$
N \geq(q+1)(q+2) r(S) \text { if } r(S) \geq 1
$$

and

$$
N \geq q^{2}+q+2 \text { if } r(S)=0 ;
$$

but the author does not know if these values are in fact the best.

For theorems of this type, the conditions that the sets $A_{j}$ (or, more generally, their closures) be connected, and that the numbers $b_{0}\left\{\mathrm{Cl}\left[\mathrm{Co}\left(A_{j}\right)\right]\right\}$ be bounded, cannot be omitted. In [4, §8] a theorem in a similar order of ideas was obtained for arbitrary connected sets in a unicoherent space; it can indeed be extended to the multicoherent case, but at the cost of requiring not only that certain intersections of the sets $A_{j}$ be nonempty, but that they have sufficiently many components. For example, the theorem for three sets becomes:
(1) If $A_{1}, A_{2}, A_{3}$, are connected subsets of $S$ such that $A_{1} \cap A_{2} \cap A_{3}=0$, and $b_{0}\left(A_{j} \cap A_{k}\right) \geq r(S)$ whenever $j \neq k$, then every two of the sets $\operatorname{Fr}\left(A_{j}\right)$ meet.

The proof of (1) is an easy consequence of [5, §7.2].
We finally remark that the present technique can be used to give a direct "elementary" proof of Theorem 1, without using mappings in $S^{1}$. However, though the basic idea (showing that the sets have the same incidences as if $S$ were a linear graph) is simple, a quite lengthy and tedious argument is needed to reduce the general theorem to the case in which the complements of the sets are of finite incidence; and the proof given in 2.2 above is considerably shorter.

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[^0]:    *When $p>1$, the factor $(-1)^{s}$ may be suppressed in the summand in (2) because the terms corresponding to odd values of $s$ vanish.

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[^1]:    presented to the American Mathematical Society at the Summer meeting of 1949 in Boulder, Colorado, under the title "Quasi-convexity and the lower semicontinuity of double integrals'.

[^2]:    ${ }^{1}$ As was remarked in $[5, \delta 6.6(3)]$, there would be no difficulty in reformulating the theorems so as to apply if complete normality were weakened to normality.

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[^3]:    ${ }^{2}$ That is, $A_{j}-A_{k}$ and $A_{k}-A_{j}$ are separated ( $1 \leq j<k \leq n$ ). (Two sets are "separated" if neither meets the closure of the other.)

[^4]:    ${ }^{3}$ Throughout this paper, "disjoint" means "pairwise disjoint'.
    ${ }^{4}$ That is, the sets $A_{j}$ and all their intersections $A_{J}$ have only finitely many components.

[^5]:    ${ }^{5}$ These hypotheses are implied by: (a) the sets $\operatorname{Fr}\left(A_{j}\right)$ are disjoint, or (b) $A_{1}, \cdots, A_{n}$ are all open, or all closed, and $\operatorname{Fr}\left(A_{j} \cap A_{k}\right) \cap \operatorname{Fr}\left(A_{j} \cup A_{k}\right)=0$ whenever $j \neq k$, or (c) $A_{1}, \cdots, A_{n}$ are all closed and $\operatorname{Fr}\left(A_{j}\right) \cap \operatorname{Fr}\left(A_{k}\right) \cap \operatorname{Fr}\left(A_{j} \cup A_{k}\right)=0(j \neq k)$, or dually, and thus also by: (d) $A_{1}, \cdots, A_{n}$ are closed and cover $S$, and no three of them have a common point. A slight relaxation of the hypotheses on the sets $A_{j}$ is possible; see 2.3 (3) below.
    ${ }^{6}$ The case $m=1$ is trivial. If equality holds in the conclusion of Theorem 1 , and both sides are finite, then the sets $A_{j}$ must in fact satisfy stronger frontier conditions; see 5.6 below.

[^6]:    ${ }^{7}$ Note that we do not require every two sets $A_{j}, A_{k}$ to cover $S$. In fact, if $n$ nonempty closed sets are such that every two of them cover $S$, then trivially all of them have a common point.

[^7]:    9 It follows (see 5.6 below) that, in the case of finite equality, we have for each component $E$ of $A \cup B$ that $\operatorname{Fr}(A \cap E) \cap \operatorname{Fr}(B \cap E)=0$. It is false, in general, that $\operatorname{Fr}(A) \cap \operatorname{Fr}(B)=0$.

[^8]:    ${ }^{13}$ A set $E$ is "simple" if $E$ and $S-E$ are both connected.
    ${ }^{14}$ It would suffice to require only that $\mathrm{Cl}\left(A_{j}\right)$ and $\mathrm{Cl}\left[\mathrm{Co}\left(A_{j}\right)\right]$ be connected ( $1 \leq j \leq n$ ).

