# Introduction to Various Aspects of Degree Theory in Banach Spaces 

E. H. Rothe

Mathematical<br>Surveys and Monographs

Volume 23.S

# Introduction to Various Aspects of Degree Theory in Banach Spaces 

E. H. Rothe

American Mathematical Society
Providence, Rhode Island

# 2000 Mathematics Subject Classification. Primary 47-XX; Secondary $54-\mathrm{XX}, 55-\mathrm{XX}, 58-\mathrm{XX}$. 

## Library of Congress Cataloging-in-Publication Data

Rothe, Erich H.
Introduction to various aspects of degree theory in Banach spaces. (Mathematical surveys and monographs, ISSN 0076-5376; no. 23)
Bibliography: p.
Includes index.

1. Degree, Topological. 2. Banach spaces. I. Title. II. Series.

QA612.R68 $1986 \quad 514^{\prime} .2 \quad$ 86-8038
ISBN 0-8218-1522-9 (alk. paper)

Copying and reprinting. Individual readers of this publication, and nonprofit libraries acting for them, are permitted to make fair use of the material, such as to copy a chapter for use in teaching or research. Permission is granted to quote brief passages from this publication in reviews, provided the customary acknowledgment of the source is given.

Republication, systematic copying, or multiple reproduction of any material in this publication is permitted only under license from the American Mathematical Society. Requests for such permission should be addressed to the Assistant to the Publisher, American Mathematical Society, P. O. Box 6248, Providence, Rhode Island 02940-6248. Requests can also be made by e-mail to reprint-permission@ams.org.
(c) Copyright 1986 by the American Mathematical Society. All rights reserved. Printed in the United States of America.
The American Mathematical Society retains all rights except those granted to the United States Government.
(a) The paper used in this book is acid-free and falls within the guidelines
established to ensure permanence and durability.
Visit the AMS home page at URL: http://www.ams.org/

$$
1098765432 \quad 060504030201
$$

## Contents

Preface ..... v
Introduction ..... 1
CHAPTER 1. Function-Analytic Preliminaries ..... 13
CHAPTER 2. The Leray-Schauder Degree for Differentiable Maps ..... 26

1. The degree for linear maps ..... 26
2. The degree if $y_{0}$ is a regular value for the map $f$ ..... 27
3. The one-dimensional case and the case of polynomial maps ..... 31
4. The degree for a not necessarily regular value $y_{0}$ ..... 39
5. Notes ..... 51
CHAPTER 3. The Leray-Schauder Degree for Not Necessarily Differentiable Maps ..... 59
6. An extension lemma ..... 59
7. An application of the extension lemma ..... 62
8. The degree theory for finite layer maps ..... 64
9. Another application of the extension lemma ..... 65
10. Two additional properties of the Leray-Schauder degree ..... 67
11. Generalized L.-S. maps ..... 69
12. Notes ..... 71
CHAPTER 4. The Poincaré-Bohl Theorem and Some of Its Applications ..... 76
13. The Poincaré-Bohl theorem and the winding number ..... 76
14. The interpretation of degree and winding number as intersection numbers ..... 83
15. Notes ..... 84
CHAPTER 5. The Product Theorem and Some of Its Consequences ..... 88
16. The product theorem ..... 88
17. The invariance of the domain ..... 97
18. The Jordan-Leray theorem ..... 103
19. Notes ..... 108
CHAPTER 6. The Finite-Dimensional Case ..... 112
20. Some elementary prerequisites ..... 112
21. The degree for mappings between finite-dimensional spaces of the same dimension ..... 117
22. Simplicial mappings ..... 122
23. On subdivisions ..... 125
24. Simplicial approximations ..... 126
25. Notes ..... 128
CHAPTER 7. On Spheres ..... 138
26. Elementary properties and orientation of spheres. Degree of mappings between spheres ..... 138
27. Properties of the degree $d\left(f, S_{1}^{n}, S_{2}^{n}\right)$ ..... 142
28. The order of the image of a sphere with respect to a point ..... 146
29. Two approximation lemmas ..... 155
30. Notes ..... 157
CHAPTER 8. Some Extension and Homotopy Theorems ..... 172
31. An extension and a homotopy theorem ..... 172
32. Two further extension theorems ..... 182
33. Notes ..... 186
CHAPTER 9. The Borsuk Theorem and Some of Its Consequences ..... 192
34. The Borsuk theorem ..... 192
35. Some consequences of the Borsuk theorem ..... 195
36. Notes ..... 198
APPENDIX A. The Linear Homotopy Theorem ..... 199
37. Motivation for the theorem and the method of proof ..... 199
38. Background material from spectral theory in a complex Banach space $Z$ ..... 200
39. The complexification $Z$ of a real Banach space $E$ ..... 204
40. On the index $j$ of linear nonsingular L.-S. maps on complex and real Banach spaces ..... 208
41. Proof of the linear homotopy theorem ..... 214
42. The multiplication theorem for the indices ..... 228
APPENDIX B. Proof of the Sard-Smale Theorem 4.4 of Chapter 2 ..... 231
References ..... 238
Index ..... 241

## Preface

As its title indicates this book is not meant to be an encyclopedic presentation of the present state of degree theory. Likewise, the list of references is not meant to be a complete bibliography. The choice of subjects treated is determined partly by the fact that the book is written from the point of view of an analyst, partly by the desire to throw light on the theory from various angles, and partly the choice is subjective.

A detailed description of the subjects treated is contained in $\S 10$ of the following introduction. At this point we mention, together with relevant references, some subject matter belonging to degree theory which is not treated in this book.
(a) The Leray-Schauder degree theory in Banach spaces may be extended to linear convex topological spaces as already noticed by Leray in [35]. A complete self-contained treatment may be found in Nagumo's 1951 paper [43].
(b) Already, in his 1912 paper [9], Brouwer established degree theory for certain finite-dimensional spaces which are not linear. For later developments in this direction and for bibliographies, see, e.g., the books by Alexandroff-Hopf [2], by Hurewicz-Wallmann [29], and by Milnor [39]. In their 1970 paper [20] Elworthy and Tromba established a degree theory in certain infinite-dimensional manifolds. See also the systematic exposition by Borisovich-Zvyagin-Sapronov [6].
(c) There are various degree theories for mappings which are not of the type treated by Leray and Schauder: a theory by Browder and Nussbaum on "intertwined" maps [11] and a theory of " $A$-proper" maps by Browder and Petryshyn [12] in which the degree is not integer-valued but a subset of the integers. See also [10]. Moreover, the method used in the Fenske paper [21] mentioned in $\S 8$ of the following introduction applies also to mappings which are " $\alpha$-contractions." Various mathematicians considered degree theory for multiple-valued maps. A survey of these generalizations and further bibliographical data may be found in the book [37] by Lloyd.
(d) Some important theorems based on degree theory (e.g., those contained in $\S \S 2-3$ of Chapter 5 and parts of Chapters 8 and 9 of this book) and some generalizations thereof can be derived by the use of cohomology theory in infinite-
dimensional spaces. See the paper by Geba and Granas [22], in particular Chapter IX, and the book by Granas [25], in particular Chapter 11.
(e) The axiomatic treatment of degree theory by Amann and Weiss [3].
(f) The "coincidence degree theory" of Mawhin [38].
(g) For supplementary reading and further references we mention books by the following authors: Krasnoselskiĭ [31], 1956, J. Cronin [14], 1964, J. T. Schwartz [51], 1965, K. Deimling [15], 1974, E. Zeidler [54], 1976, F. Browder [10], 1976, G. Eisenack-C. Fenske [19], 1978, N. G. Lloyd [37], 1978. We also mention the paper [52], 1980 by H. W. Siegbert which contains an interesting exposition of the history of finite-dimensional degree theory from Gauss via Kronecker to Brouwer.
(h) Ever since it was established by Leray and Schauder [36], degree theory in Banach spaces has been an important tool for the treatment of boundary value problems (including periodicity problems) for ordinary and partial differential equations, for integral equations, and for eigenvalue and bifurcation problems. These applications to analysis are not treated in the present book, but the reader will find a number of them discussed in many of the books quoted in part (g) of this preface. See, e.g., Chapters III-VI of [31], Chapters II and IV of [14], and Chapter 9 of [37]. The Cronin book [14] discusses, in particular, papers up to 1962 using the "Cesari method." For the role of degree theory in the further development of this method, now usually referred to as the "Alternative Method," the reader may consult the bibliography on pp. 232-234 in Nonlinear Phenomena in Mathematical Science, ed. V. Lakshmikantham, Academic Press, 1982.

We finish this preface with a word on the organization of this book. In addition to the introduction the book consists of nine chapters. Each chapter (except Chapter 1) is divided into sections and each section into subsections. (The notation " 1.2 " in Chapter 3 means subsection 1.2 in that chapter; "(1.2)" refers to a formula in that chapter. However $\S 1.2$ refers to $\S 2$ in Chapter 1.) Each chapter is followed by notes containing some of the proofs and historical remarks. Finally, there are two Appendixes A and B.

My thanks go to my colleagues and friends Professors Lamberto Cesari, Charles Dolph, and R. Kannan for many encouraging conversations.

My thanks also go to Mrs. Wanita Rasey who, with great patience, converted an often not easily decipherable handwritten script into a readable typescript.

I am obliged to the editorial board of the Mathematical Surveys and Monographs for reviewing the typescript and to the editorial staff of the American Mathematical Society, in particular to Ms. Mary C. Lane, Lenore C. Stanoch, and Holly Pappas for their cooperation and labor in transforming the manuscript into a book.

## Appendix A

## The Linear Homotopy Theorem

## §1. Motivation for the theorem and the method of proof

1.1. Let $E$ be a Banach space, and let

$$
\begin{equation*}
l(x)=x-L(x) \tag{1.1}
\end{equation*}
$$

be a nonsingular L.-S. map, $E \rightarrow E$. We want to define an index $j(l)$ which indicates how often $E$ is covered by $l(E)$. Since $l$ is a nonsingular L.-S. map, it maps, according to $\S \S 1.11$ and $1.12, E$ onto $E$ in a one-to-one fashion. Therefore we should define $j(l)$ as a number of absolute value 1 . In order to decide for which $l$ to define $j(l)$ as +1 and for which as -1 , we consider the finite-dimensional case.

Let $E=E^{n}$ be an oriented space of finite dimension $n$ (see subsections 1.21 and 1.22 in Chapter 6 for the concept of orientation). It is then natural to define $j(l)$ as +1 if $l$ preserves the orientation (as does, e.g., the identity map $I$ ) and as -1 otherwise. By the sections just quoted, this is equivalent to defining

$$
j(l)= \begin{cases}+1 & \text { if } \operatorname{det} l>0  \tag{1.2}\\ -1 & \text { if } \operatorname{det} l<0\end{cases}
$$

where $\operatorname{det} l$ denotes the determinant of $l$. (Note that $\operatorname{det} l \neq 0$ since $l$ is supposed to be nonsingular.) Now definition (1.2) in the form stated cannot be used in general Banach spaces since it depends on determinants; however, it can be reformulated in a way to become meaningful for Banach spaces. This is possible on account of the following facts (a) and (b) which are well known in linear algebra, being a consequence of the existence of the Jordan normal form for $n \times n$ matrices (see, e.g., $[26, \S 58$ and $\S 77]$ ): (a) if $L$ has real eigenvalues greater than 1 , say $\lambda_{1}>\lambda_{2}>\cdots>\lambda_{r}>1$, and if for $\rho=1, \ldots, r, \mu_{\rho}=\mu\left(\lambda_{\rho}\right)$ is the generalized multiplicity of $\lambda_{\rho}$, i.e., the dimension of the subspace $E^{\rho}$ of $E^{n}$ which consists of those $x \in E^{n}$ for which

$$
\left(\lambda_{\rho} I-L\right)^{m} x=\theta
$$

for some positive integer $m$, then

$$
\begin{equation*}
\operatorname{sign} \operatorname{det} l=(-1)^{\sum_{\rho=1}^{r} \mu_{\rho}} ; \tag{1.3}
\end{equation*}
$$

(b) if $L$ has no real eigenvalues greater than 1 , then

$$
\begin{equation*}
\text { sign } \operatorname{det} l=+1 \tag{1.4}
\end{equation*}
$$

It follows from (1.3) and (1.4) that definition (1.2) is equivalent to

$$
j(l)= \begin{cases}(-1)^{\sum_{\rho=1}^{r} \mu_{\rho}} & \text { in case (a) }  \tag{1.5}\\ +1 & \text { in case (b) }\end{cases}
$$

As is well known and as will be seen in the following sections, the right member of (1.5) makes sense for an arbitrary Banach space $E$ if $l$ is a nonsingular linear L.-S. map $E \rightarrow E$. We note that Definition 1.6 in Chapter 2 appears as a theorem in the original Leray-Schauder theory (see [36, II. 11]).

In $\S 2$ of this appendix we recall background material from the spectral theory of linear operators $M$, referring for proofs mainly to the presentation given in [18, Chapter VII]. Since in general the spectrum of $M$ contains complex numbers, we deal here with a complex Banach space $Z$. In order to apply the theory to the given real Banach space $E$, we have to "complexify" $E$, i.e., embed it into a complex Banach space $Z$. This is done in $\S 3$. $\S 4$ discusses the index $j(l)$ in $E$ as defined by (1.5) and also its definition for the complexification $Z$ of $E$. Finally, in $\S 5$, the linear homotopy theorem (cf. subsection 1.3 in Chapter 2) is proved.

## §2. Background material from the spectral theory in a complex Banach space $Z$

2.1. Definition. Let $Z$ be a complex Banach space, let $M$ be a linear bounded operator on $Z$, and let $I$ denote the identity map on $Z$. Then the set of complex numbers $\lambda$ for which the inverse $R_{\lambda}(M)$ of $\lambda I-M$ exists as a bounded operator $Z \rightarrow Z$ is called the resolvent set of $M$ and will be denoted by $\rho(M)$. The complement $\sigma(M)$ of $\rho(M)$ in the $\lambda$-plane is called the spectrum of M. $\sigma(M)$ is closed and bounded. A subset $\sigma_{0}(M)$ of $\sigma(M)$ is called a spectral set of $M$ if it is open and closed with respect to $\sigma(M)[18, \mathrm{pp} .566,567,572]$.
2.2. For fixed $M$ the domain of $R_{\lambda}(M)$ as a function of $\lambda$ is the set $\rho(M)$, and its range is a subset of the family of bounded linear operators on $Z$. This family is a Banach space with the linear operations defined in the obvious way and with the norm defined [18, pp. 475, 477] by

$$
\begin{equation*}
\|M\|=\sup _{\|z\| \leq 1}\|M(z)\| \tag{2.1}
\end{equation*}
$$

This Banach space will be denoted by $Z_{1}$, and for $M$ fixed we consider $R_{\lambda}(M)$ as a map $\rho(M) \rightarrow Z_{1}$. The topology of $Z_{1}$ induced by the norm (2.1) is called the uniform operator topology.
2.3. An example for a spectral subset of the spectrum $\sigma(M)$ is $\sigma(M)$ itself, and so is every isolated point of $\sigma(M)$.
2.4. Lemma. $\rho(M)$ is open, and $\sigma(M)$ is a nonempty closed bounded set. $R_{\lambda}(M)$ as a function of $\lambda$ is analytic at each $\lambda \in \rho(M)$, i.e., at such $\lambda$ the
derivative of $R_{\lambda}(M)$ with respect to $\lambda$ exists. Obviously

$$
\sum_{n=0}^{\infty}\|M\|^{n}|\lambda|^{-(n+1)}
$$

converges if

$$
\begin{equation*}
|\lambda|>\|M\| . \tag{2.2}
\end{equation*}
$$

Moreover for such $\lambda$

$$
\begin{equation*}
R_{\lambda}(M)=\sum_{n=0}^{\infty} \frac{M^{n}}{\lambda^{n+1}} . \tag{2.3}
\end{equation*}
$$

For a proof of this lemma see [18, pp. 566,567].
2.5. Lemma. Let $\Delta$ be a closed not necessarily bounded subset of $\rho(M)$. Then $R_{\lambda}(M)$ is bounded on $\Delta$.

Proof. On the bounded closed set $\{\lambda||\lambda| \leq\|M\|+1\}$ the boundedness of $R_{\lambda}(M)$ follows from its analyticity. But on the set $\{\lambda||\lambda|>\|M\|+1\}$ it follows from (2.3).
2.6. Definition. $\mathcal{F}(M)$ denotes the family of all complex-valued functions $f$ which are defined and analytic in some open neighborhood $U$ of $\sigma(M)$. ( $U$ may depend on $f$.)
2.7. Let $\sigma_{0}$ be a spectral set of $M$, and let $V_{0}$ be an open set in the complex $\lambda$-plane containing $\sigma_{0}$ but no point of $\sigma-\sigma_{0}$ where $\sigma=\sigma(M)$. Let $f \in \mathcal{F}(M)$ and let $U_{f}$ be the domain of definition of $f$. Let $\Gamma_{0}$ and $\Gamma_{1}$ be two rectifiable Jordan curves lying in $V_{0} \cap U_{f}$ oriented in the counterclockwise sense and such that $\sigma_{0}$ lies in the open bounded sets whose boundaries are $\Gamma_{0}$ and $\Gamma_{1}$. Then

$$
\begin{equation*}
\int_{\Gamma_{0}} f(\mu) R_{\mu}(M) d \mu=\int_{\Gamma_{1}} f(\mu) R_{\mu}(M) d \mu \tag{2.4}
\end{equation*}
$$

We note that the integrands in (2.4) are operator-valued. For the definition of such integrals and for the proof of the "Cauchy theorem" (2.4), we refer the reader to [18, pp. 224-227].
2.8. Definition. For $f \in \mathcal{F}(M)$ we define $f(M)$ by setting

$$
\begin{equation*}
f(M)=\frac{1}{2 \pi i} \int_{\Gamma} f(\mu) R_{\mu}(M) d \mu \tag{2.5}
\end{equation*}
$$

where $\Gamma$ is a rectifiable Jordan curve of the following properties: (i) $\sigma(M)$ lies in the bounded open domain whose boundary is $\Gamma$; (ii) $\Gamma$ is contained in the domain $U_{f}$ of definition for $f$; (iii) $\Gamma$ is oriented in the counterclockwise sense. The uniqueness of Definition 2.5 follows from Lemma 2.7.
2.9. Lemma. Let $\Gamma$ be a rectifiable Jordan curve satisfying conditions (i), (ii) and (iii) of subsection 2.8. Then

$$
\begin{equation*}
M=\frac{1}{2 \pi i} \int_{\Gamma} \mu R_{\mu}(M) d \mu \tag{2.6}
\end{equation*}
$$

$$
\begin{equation*}
I=\frac{1}{2 \pi i} \int_{\Gamma} R_{\mu}(M) d \mu \tag{2.7}
\end{equation*}
$$

If, moreover, $f$ and $g$ are elements of $\mathcal{F}(M)$ with domains $U_{f}$ and $U_{g}$ resp. and if, in addition to the above requirements, $\Gamma \in U_{f} \cap U_{g}$, then

$$
\begin{equation*}
f(M) \cdot g(M)=\frac{1}{2 \pi i} \int_{\Gamma} f(\mu) g(\mu) R_{\mu}(M) d \mu \tag{2.8}
\end{equation*}
$$

For the proof we refer the reader to the proof of Theorem 10, p. 468 in [18], which contains the assertions of the present lemma.
2.10. Definition. A bounded linear map $P: Z \rightarrow Z$ is called a projection if

$$
\begin{equation*}
P^{2}=P \tag{2.9}
\end{equation*}
$$

2.11. Lemma. Let $\sigma_{0}$ be a spectral set of $M$. Let $\Gamma_{0}$ be as in Lemma 2.7. Then the operator $P_{\sigma_{0}}$ defined by

$$
\begin{equation*}
P_{\sigma_{0}}=\frac{1}{2 \pi i} \int_{\Gamma_{0}} R_{\mu}(M) d \mu \tag{2.10}
\end{equation*}
$$

is a projection.
Proof. Let $\sigma_{0}^{*}=\sigma-\sigma_{0}$. By the definition of spectral sets, there exist bounded open sets $V_{0}, V_{0}^{*}, U_{0}, U_{0}^{*}$ such that $\sigma_{0} \subset V_{0} \subset \bar{V}_{0} \subset U_{0}, \sigma_{0}^{*} \subset V_{0}^{*} \subset$ $\bar{V}_{0}^{*} \subset U_{0}^{*}$ with $\bar{U}_{0}$ and $\bar{U}_{0}^{*}$ disjoint and such that $\Gamma_{0}=\partial V_{0}$ and $\Gamma^{*}=\partial V^{*}$ are rectifiable Jordan curves. Let $\Gamma=\Gamma_{0} \cup \Gamma^{*}$.

Then (2.8) holds if we set

$$
f(\mu)=g(\mu)= \begin{cases}1, & \mu \in U_{0}  \tag{2.11}\\ 0, & \mu \in U_{0}^{*}\end{cases}
$$

But with this choice of $f$ and $g$ we see from (2.10) that the right member of (2.8) equals $P_{\sigma_{0}}$ since $f(\mu)=g(\mu)=0$ for $\mu \in \Gamma_{0}^{*}$, and taking account also of (2.5), we see that $f(M)=g(M)=P_{\sigma_{0}}$. Thus the left member of (2.8) equals $P_{\sigma_{0}}^{2}$.
2.12. Lemma. Let $\sigma_{0}$ and $\sigma_{1}$ be two disjoint spectral sets for $M$. Let $\Gamma_{0}$ and $P_{\sigma_{0}}$ be defined as in subsection 2.11, and let $\Gamma_{1}$ and $P_{\sigma_{1}}$ be defined correspondingly with respect to $\sigma_{1}$. Then

$$
\begin{equation*}
P_{\sigma_{0}} \cdot P_{\sigma_{1}}=P_{\sigma_{1}} \cdot P_{\sigma_{0}}=\theta_{1} \tag{2.12}
\end{equation*}
$$

where $\theta_{1} x=\theta$ for every $x \in E$.
Proof. Since $\sigma_{0}, \sigma_{1}$, and $\sigma_{2}=\sigma-\left(\sigma_{1} \cup \sigma_{2}\right)$ are closed bounded sets, there exist for $i=0,1,2$ bounded open sets $V_{i}, U_{i}$ such that $\sigma_{i} \subset V_{i} \subset \bar{V}_{i} \subset U_{i}$ with the sets $\bar{U}_{1}, \bar{U}_{2}, \bar{U}_{3}$ being pairwise disjoint and such that $\Gamma_{i}=\partial V_{i}$ is a rectifiable Jordan curve which we assume to be oriented in the counterclockwise sense. If we set

$$
f(\mu)=\left\{\begin{array}{ll}
1 & \text { for } \mu \in U_{0}, \\
0 & \text { for } \mu \in U_{1} \cup U_{2},
\end{array} \quad g(\mu)= \begin{cases}1 & \text { for } \mu \in U_{1} \\
0 & \text { for } \mu \in U_{0} \cup U_{2}\end{cases}\right.
$$

then with $\Gamma=\Gamma_{1} \cup \Gamma_{2} \cup \Gamma_{3}$

$$
\begin{aligned}
P_{\sigma_{0}} & =\frac{1}{2 \pi i} \int_{\Gamma} f(\mu) R_{\mu}(M) d \mu=f(M) \\
P_{\sigma_{1}} & =\frac{1}{2 \pi i} \int_{\Gamma} g(\mu) R_{\mu}(M) d \mu=g(M)
\end{aligned}
$$

and application of (2.8) yields the assertion (2.12) since $f(\mu) \cdot g(\mu)=0$ for all $\mu \in \Gamma$.
2.13. Lemma. Let $\sigma_{0}$ be a spectral set of $M$, let $P_{\sigma_{0}}$ be the projection (2.10), and let $Z_{0}$ be the range of $P_{\sigma_{0}}$, i.e., $Z_{0}=P_{\sigma_{0}} Z$. Then $M Z_{0} \subset Z_{0}$ and $\sigma_{0}$ is the spectrum of the restriction of $M$ to $Z_{0}$.

For a proof see [18, pp. 574, 575].
2.14. Lemma. The restriction $M_{\sigma_{0}}$ of $M$ to $Z_{0}=P_{\sigma_{0}} Z$ is given by

$$
\begin{equation*}
M_{\sigma_{0}}=\int_{\Gamma_{0}} \mu R_{\mu}(M) d \mu \tag{2.13}
\end{equation*}
$$

where $\Gamma_{0}$ is as in subsection 2.7.
Proof. Let $\Gamma$ be as in Lemma 2.9, and let $V_{0}$ and $V$ be the bounded open sets whose boundaries are $\Gamma_{0}$ and $\Gamma$ resp. Then $M$ is given by (2.6), i.e., by (2.5) with $f(\mu)=\mu$. Now $z=P_{\sigma_{0}}(z)$ if and only if $z \in Z_{0}=P_{\sigma_{0}} Z$. Therefore $M(z)=M P_{\sigma_{0}}(z)$ for $z \in Z_{0}$. But $P_{\sigma_{0}}$ is given by (2.10) or what is the same

$$
\begin{equation*}
P_{\sigma_{0}}=\frac{1}{2 \pi i} \int_{\Gamma} g(\mu) R_{\mu}(M) d \mu \tag{2.14}
\end{equation*}
$$

with $g(\mu)$ as in (2.11). The assertion (2.13) follows now from (2.5) and (2.14) on application of (2.8).
2.15. So far we assumed of the linear operator $M$ only boundedness or what is the same continuity. In the remainder of this section we require complete continuity, i.e., we make the additional assumption that for any bounded set $\beta \subset E$ the closure of $M(\beta)$ is compact. The definitions and assertions stated below are classical.

Proofs may be found, e.g., in [18, p. 577ff.].
2.16. The spectrum $\sigma(M)$ of a completely continuous linear operator $M$ on $E$ is at most countable. If $\sigma(M)$ contains infinitely many points, then 0 is the only accumulation point of $\sigma(M)$. It follows that each $\lambda_{0} \in \sigma(M)$ which is different from 0 is a spectral set $\sigma_{0}$ of $M$. For the projection $P_{\sigma_{0}}$ defined by (2.10) we write:

$$
\begin{equation*}
P_{\lambda_{0}}=\frac{1}{2 \pi i} \int_{\Gamma_{0}} R_{\mu}(M) d \mu \tag{2.15}
\end{equation*}
$$

For $\Gamma_{0}$ we may and will take a circle with center $\lambda_{0}$ and radius smaller than the distance of $\lambda_{0}$ from the rest of $\sigma(M) . \Gamma_{0}$ is supposed to be oriented in the counterclockwise sense. We recall that the points of $\sigma(M)$ are also referred to as eigenvalues of $M$.
2.17. Let $\lambda_{0} \in \sigma(M)$ and $\lambda_{0} \neq 0$. Then

$$
\begin{equation*}
P_{\lambda_{0}} Z=\left\{z \in Z \mid\left(\lambda_{0} I-M\right)^{n}=0\right\} \tag{2.16}
\end{equation*}
$$

for some positive integer $n$. The dimension $\nu\left(\lambda_{0}\right)$ of the linear subspace $P_{\lambda_{0}} Z$ of $Z$ is finite and is called the generalized multiplicity of $\lambda_{0}$ as an eigenvalue of $M$.
2.18. Let $\lambda_{0} \in \sigma(M)$ and $\lambda_{0} \neq 0$. Then the equation

$$
\begin{equation*}
\lambda_{0} z-M(z)=\theta \tag{2.17}
\end{equation*}
$$

has nontrivial solutions (i.e., solutions $z \neq 0$ ) called eigenelements of $M$ to the eigenvalue $\lambda_{0}$. These eigenelements obviously form a linear subspace of $E$. The dimension $r\left(\lambda_{0}\right)$ of this subspace satisfies the inequality $1 \leq r\left(\lambda_{0}\right) \leq \nu\left(\lambda_{0}\right)$ and is called the multiplicity of $\lambda_{0}$ as an eigenvalue of $M$.
2.19. Lemma. Let $M$ be a completely continuous linear map $Z \rightarrow Z$ and suppose that the linear map $m=I-M$ mapping $Z$ onto $Z$ is one-to-one. Then the inverse $n=m^{-1}$ exists and is of the form $m(\xi)=\xi-N(\xi)$, where $N$ is completely continuous on $Z$.

Proof. By assumption $m$ is a linear, continuous, and therefore bounded, one-to-one map of $Z$ onto $Z$. By a well-known theorem (see, e.g., $[18$, p. 57$]$ ) these properties imply that the linear map $n=m^{-1}$ is bounded. Now if $\xi=$ $z-M(z)$, then $n(\xi)=z=\xi+M(z)=\xi+M(n(\xi))$. Since $n$ is bounded, i.e., continuous, and $M$ is completely continuous, the map $N(\xi)=-M(n(\xi))$ is completely continuous.

## §3. The complexification $Z$ of a real Banach space $E$

3.1. The construction of $Z$ from $E$ is analogous to the construction of complex numbers as ordered pairs of real numbers: the elements $z$ of $Z$ are ordered pairs $(x, y)$ of points of the real Banach space $E$. If $z_{i}=\left(x_{i}, y_{i}\right), i=1,2$, are two elements of $Z$, then their sum is defined by $z_{1}+z_{2}=\left(x_{1}+x_{2}, y_{1}+y_{2}\right)$, and if $c=\alpha+i \beta, \alpha, \beta$ real, is a complex number, then $c z_{1}$ is defined by

$$
\begin{equation*}
c z_{1}=\left(\alpha x_{1}-\beta y_{1}, \alpha y_{1}+\beta x_{1}\right) \tag{3.1}
\end{equation*}
$$

Finally if as usual $\|\cdot\|$ denotes the norm in $E$, we set for the point $z=(x, y) \in Z$

$$
\begin{equation*}
\|z\|_{1}=\sup _{\phi}\|x \cos \phi+y \sin \phi\| \tag{3.2}
\end{equation*}
$$

3.2. In the case that $E$ is the real line, then $Z$ is the complex plane. The reader may verify that then $\|z\|_{1}$ is the absolute value of the complex number $z$.
3.3. The number $\|z\|_{1}$ defined by (3.2) is a norm, i.e., for points $z, z_{1}, z_{2}$ in $Z$ and complex number $c$

$$
\begin{equation*}
\|z\|_{1} \geq 0 \tag{3.3}
\end{equation*}
$$

with the equality holding if and only if $z=\theta$, the zero element of $Z$, and

$$
\begin{align*}
\left\|z_{1}+z_{2}\right\| & \leq\left\|z_{1}\right\|+\left\|z_{2}\right\|  \tag{3.4}\\
\|c z\|_{1} & =|c| \cdot\|z\| \tag{3.5}
\end{align*}
$$

3.4. With the norm $\|z\|_{1}$ the linear space $Z$ is complete and thus a Banach space.

The verification of the assertions contained in subsections 3.3 and 3.4 is of course based on the corresponding properties of the Banach space $E$ of which $Z$ is the complexification. We leave this verification to the reader.
3.5. It is easily verified from our definitions that $\left\|\left(x, \theta_{1}\right)\right\|_{1}=\|x\|$. Thus the linear one-to-one map $E$ into $Z$ given by

$$
\begin{equation*}
x \rightarrow(x, \theta) \tag{3.6}
\end{equation*}
$$

is norm preserving. Identifying $x \in E$ with $(x, \theta) \in Z$, we consider $E$ as a subset of $Z$. This identification also allows us to write $\|z\|$ instead of $\|z\|_{1}$ without ambiguity.
3.6. From the multiplication rule (2.1) we see that $(\theta, y)=i(y, \theta)$ for any $y \in$ $E$ where as usual $i=(0,1)$, the imaginary unit of the complex plane. Therefore $z=(x, y)=(x, \theta)+(\theta, y)=(x, \theta)+i(y, \theta)$, and by our above identification we see that every $z \in Z$ may be written uniquely in the form

$$
\begin{equation*}
z=x+i y, \quad x \in E, y \in E \tag{3.7}
\end{equation*}
$$

3.7. For later use we note the inequality

$$
\begin{equation*}
\frac{1}{2}(\|x\|+\|y\|) \leq\|z\| \leq\|x\|+\|y\| \quad \text { for } z=x+i y \tag{3.8}
\end{equation*}
$$

The right part of this inequality is a restatement of the triangle inequality (3.4). To prove the left part we note that $\|x\|=\|x \cos 0+y \sin 0\| \leq$ $\sup _{\phi}\|x \cos \phi+y \sin \phi\|=\|z\|$ by (3.2). In the corresponding way we see that $\|y\| \leq\|z\|$. Thus $\|x\|+\|y\| \leq 2\|z\|$.
3.8. Definition. Let $L$ be a linear bounded operator $E \rightarrow E$. Then the map $M: Z \rightarrow Z$ defined for $z=x+i y$ by

$$
\begin{equation*}
M(z)=L(x)+i L(y) \tag{3.9}
\end{equation*}
$$

is called the complex extension of $L$.
3.9. Lemma. (i) $M$ is linear; (ii) $M$ is bounded; (iii) $M$ is completely continuous if and only if $L$ is completely continuous; (iv) if $L$ is continuous, then $m=I-M$ is nonsingular if and only if $l=I-L$ is not singular ( $I$ denotes the identity map in the respective space); (v) $m(x)=\theta$ has a solution $z \neq \theta$ if and only if $l(x)=\theta$ has a solution $x \neq \theta$; (vi) $m$ is not singular (i.e., $m(Z)=Z$ ) if and only if $m$ is one-to-one, i.e., if $z=\theta$ is the only root of $m(z)=\theta$.

Proof. The proof of (i) consists of an obvious verification.
Proof of (ii). Let $\mu$ be a bound for $L$. Then by (3.9), (3.4), and (3.8)

$$
\|M(z)\| \leq\|L(x)\|+\|L(y)\| \leq \mu(\|x\|+\|y\|) \leq 2 \mu\|z\| .
$$

Proof of (iii). Suppose $L$ is completely continuous. Then, by (ii), $M$ is continuous and it remains to show that if $z_{n}=x_{n}+i y_{n}$ where $x_{n}$ and $y_{n}$ are
bounded sequences in $E$, then the $M\left(z_{n_{i}}\right)$ converge for some subsequence $z_{n_{i}}$ of the $z_{n}$. Now by (3.8) the boundedness of the $z_{n}$ implies the boundedness of the $x_{n}$ and $y_{n}$. It therefore follows from the complete continuity of $L$ that for some subsequence $n_{i}$ of the integers the $L\left(x_{n_{i}}\right)$ and $L\left(y_{n_{i}}\right)$ converge. But then for $z_{n_{i}}=x_{n_{i}}+i y_{n_{i}}$ the sequence $M\left(z_{n_{i}}\right)=L\left(x_{n_{i}}\right)+i L\left(y_{n_{i}}\right)$ converges. The converse is obvious since $L$ is a restriction of $M$ to $E$.

Proof of (iv): If $z=x+i y, \varsigma=\xi+i \eta$, with $x, y, \xi, \eta$ elements of $E$, then the equation $m(z)=\varsigma$ is equivalent to the couple of equations $l(x)=\xi, l(y)=\eta$. This obviously implies that $l$ maps $E$ onto $E$ if and only if $m$ maps $Z$ onto $Z$. We obtain a proof of assertion (v) if in the proof of (iv) we set $\zeta=\xi=\eta=\theta$. Finally assertion (vi) follows from assertions (iv) and (v) in conjunction with part (iv) of Lemma 12 in Chapter 1, the latter being valid also for complex Banach spaces.
3.10. Definition. With $E$ considered as subset of $Z$ (cf. subsection 3.5), the elements of $E$ are called the real elements of $Z$. If $z=x+i y$ with $x, y$ real, the element $\bar{z}=x-i y$ is called conjugate to $z$ and the map $z \rightarrow \bar{z}$ is called a conjugation. A subspace $Z_{1}$ of $Z$ is called invariant under conjugation if $z \in Z_{1}$ implies that $\bar{z} \in Z_{1}$.
3.11. Lemma. Let $M$ be a bounded linear map $Z \rightarrow Z$. Then $M$ is the complex extension of a linear bounded map $L: E \rightarrow E$ if and only if for all $z \in Z$

$$
\begin{equation*}
M(\bar{z})=\overline{M(z)} \tag{3.10}
\end{equation*}
$$

Proof. The necessity is obvious from Definition 3.8. Then let $M$ be a linear bounded map $Z \rightarrow Z$ satisfying (3.10). Now for each $x \in E, M(x)$ is an element of $Z$. Therefore

$$
\begin{equation*}
M(x)=L_{1}(x)+i L_{2}(x) \tag{3.11}
\end{equation*}
$$

where $L_{1}(x)$ and $L_{2}(x)$ are elements of $E$.
It will be sufficient to show that $L_{2}(x)=\theta$, for then $M$ will be the complex extension of $L_{1}$. Now by (3.11) for $z=x+i y$

$$
\begin{equation*}
M(z)=M(x)+i M(y)=L_{1}(x)-L_{2}(y)+i\left(L_{2}(x)+L_{1}(y)\right) \tag{3.12}
\end{equation*}
$$

This holds for all $z \in Z$. Therefore

$$
\begin{equation*}
M(\bar{z})=L_{1}(x)+L_{2}(y)+i\left(L_{2}(x)-L_{1}(y)\right) \tag{3.13}
\end{equation*}
$$

On the other hand, we see from (3.12) that

$$
\begin{equation*}
\overline{M(z)}=L_{1}(x)-L_{2}(y)-i\left(L_{2}(x)+L_{1}(y)\right) \tag{3.14}
\end{equation*}
$$

By assumption (3.10) the left members of (3.13) and (3.14) are equal. Comparing the imaginary parts of the latter equalities, we see that $L_{2}(x)=-L_{2}(x)$, i.e., $L_{2}(x)=\theta$ as we wanted to prove.
3.12. Lemma. Let $Z_{1}$ be a linear subspace of $Z$ and let $E_{1}$ be the set of real elements in $Z_{1}$. It is asserted:
(i) if $Z_{1}$ is invariant under conjugation, then for each $z=x+i y \in Z_{1}$ the elements $x$ and $y$ belong to $E_{1}$;
(ii) if, besides satisfying the assumption of (i), $Z_{1}$ is finite dimensional, then there exists a base for $Z_{1}$ which consists of elements of $E_{1}$;
(iii) a finite-dimensional subspace $Z_{1}$ of $Z$ is invariant under conjugation if and only if there exists a projection $P: Z$ onto $Z_{1}$, such that

$$
\begin{equation*}
P(\bar{z})=\overline{P(z)} \text { for all } z \in Z \tag{3.15}
\end{equation*}
$$

PROOF. (i) If $z=x+i y \in Z_{1}\left(x, y\right.$ real), then by assumption $\bar{z}=x-i y \in Z_{1}$. Consequently $x=(z+\bar{z}) / 2$ and $y=(z-\bar{z}) / 2 i$ lie in $Z_{1} \cap E=E_{1}$.

Proof of (ii). $E_{1}$ is obviously a linear subspace of the real space $E$.
Moreover, if $b_{1}, b_{2}, \ldots, b_{r}$ are linearly independent elements of this subspace, they are also linearly independent as elements of $Z_{1}$. For if

$$
\sum_{j=1}^{r} c_{j} b_{j}=\theta
$$

with $c_{j}=\alpha_{j}+i \beta_{j}, a_{j}, \beta_{j}$ real, then

$$
\sum_{j=1}^{r} \alpha_{j} b_{j}+i \sum_{j=1}^{r} \beta_{j} b_{j}=\theta
$$

and thus

$$
\sum_{j=1}^{r} \alpha_{j} b_{j}=\sum_{j=1}^{r} \beta_{j} b_{j}=\theta
$$

Therefore $\alpha_{j}=\beta_{j}=0, j=1,2, \ldots, r$, on account of the independence of the $b_{j}$ as elements of $E_{1}$. Thus $c_{j}=\alpha_{j}+i \beta_{j}=0$. This proves the independence of the $b_{j}$ as elements of $Z_{1}$. It follows that the subspace $E_{1}$ of $E$ is finite dimensional since $Z_{1}$ is finite dimensional. Then let $b_{1}, \ldots, b_{n}$ be a base for $E_{1}$. Since the $b_{j}$ are independent also as elements of $Z_{1}$, it remains to show that they span $Z_{1}$. Now let $z=x+i y \in Z_{1}$. Then there exist real numbers $\alpha_{j}, \beta_{j}$ such that

$$
\begin{equation*}
x=\sum_{j=1}^{n} \alpha_{j} b_{j}, \quad y=\sum_{j=1}^{n} \beta_{j} b_{j} \tag{3.16}
\end{equation*}
$$

and thus

$$
z=x+i y=\sum_{j=1}^{n}\left(\alpha_{j}+i \beta_{j}\right) b_{j}
$$

Proof of (iii). The coefficients $\alpha_{j}$ in the first sum in (3.16) are obviously linear in $x$ for $x$ in the finite-dimensional subspace $E_{1}$ of $E$. They are therefore also continuous (see, e.g., [18, p. 245]). Consequently, by the Hahn-Banach theorem, these bounded linear functions $\alpha_{j}=\alpha_{j}(x)$ can be extended to bounded
linear functions $A_{j}(x)$ defined for all $x \in E$. We now define for $j=1,2, \ldots, n$ continuous linear functionals $C_{j}$ on $Z$ by setting for $z=x+i y$

$$
\begin{equation*}
C_{j}(z)=A_{j}(x)+i A_{j}(y) \tag{3.17}
\end{equation*}
$$

Then

$$
\begin{equation*}
C_{j}(\bar{z})=\overline{C_{j}(z)} \tag{3.18}
\end{equation*}
$$

since the $A_{j}(x)$ are real-valued. We now set

$$
\begin{equation*}
P(z)=\sum_{j=1}^{n} C_{j}(z) b_{j} \tag{3.19}
\end{equation*}
$$

where the $b_{j}$ form a real base for $Z_{1}$. We then see from (3.18) that (3.15) is satisfied. We now show that $P$ is a projection, i.e., that (2.9) is satisfied. Since the $b_{j}$ are real, we see from (3.19) and (3.17) that

$$
\begin{equation*}
P^{2}(z)=P(P(z))=\sum_{j=1}^{n} b_{j} C_{j}\left(\sum_{l=1}^{n} C_{l}(z) b_{l}\right)=\sum_{j=1}^{n} b_{j} \sum_{l=1}^{n} C_{l}(z) C_{j}\left(b_{l}\right) \tag{3.20}
\end{equation*}
$$

Since $b_{l} \in E_{1}, C_{j}\left(b_{l}\right)=A_{j}\left(b_{l}\right)=\alpha_{j}\left(b_{l}\right)$. But $\alpha_{j}\left(b_{l}\right)=0$ if $j \neq l$, and 1 for $j=l$ as is seen from (3.16) with $x=b_{l}$. Therefore

$$
C_{j}\left(b_{l}\right)= \begin{cases}1 & \text { if } j=l \\ 0 & \text { if } j \neq l\end{cases}
$$

and we see from (3.20) and (3.19) that

$$
P^{2}(z)=\sum_{j=1}^{n} b_{j} C_{j}(z)=P(z)
$$

This finishes the proof that the invariance of $Z_{1}$ under conjugation implies the existence of a projection $P$ satisfying (3.15).

Conversely if such projection $P$ exists, then for $z \in Z_{1}$ we see that $z=P(z)$, and therefore by (3.15) that $\bar{z}=\overline{P(z)}=P(\bar{z})$ which shows that $\bar{z} \in Z_{1}$.

## §4. On the index $j$ of linear nonsingular L.-S. maps on complex and real Banach spaces

4.1. Definition. Let $Z$ be a complex Banach space. Let

$$
\begin{equation*}
m(z)=z-M(z) \tag{4.1}
\end{equation*}
$$

be a nonsingular L.-S. map $Z \rightarrow Z$. Then with $R(\lambda)$ denoting the real part of the complex number $\lambda$, the index $j(m)$ of $m$ is defined as follows: if $M$ has no eigenvalues $\lambda$ with $\mathcal{R}(\lambda)>1$ we set $j(m)=1$. If $M$ does have eigenalues $\lambda$ with $\mathcal{R}(\lambda) \geq 1$, we denote them by $\lambda_{1}, \lambda_{2}, \ldots, \lambda_{s}$ and define

$$
\begin{equation*}
j(m)=(-1)^{\sum_{i=1}^{s} \nu_{i}} \tag{4.2}
\end{equation*}
$$

where $\nu_{i}=\nu\left(\lambda_{i}\right)$ is the generalized multiplicity of $\lambda_{i}$ as an eigenvalue of $M$ (see subsection 2.17 for the definition of "generalized multiplicity," and see $\S 1$
for a motivation for the definition of $j(m)$ ). Note that by subsection 2.16 the number of eigenvalues $\lambda$ of $M$ with $\mathcal{R}(\lambda) \geq 1$ is finite and by subsection 2.17 the generalized multiplicity by $\nu(\lambda)$ of an eigenvalue $\lambda$ of $M$ is finite.
4.2. Definition. Let $E$ be a real Banach space, and let

$$
\begin{equation*}
l(x)=x-L(x) \tag{4.3}
\end{equation*}
$$

be a nonsingular L.-S. map $E \rightarrow E$. An eigenvalue $\lambda_{0}$ of $L$ is then a real number such that for some point $x \in E$ different from $\theta$

$$
\begin{equation*}
L(x)=\lambda_{0} x \tag{4.4}
\end{equation*}
$$

If $Z$ is the complexification of $E$ and $M$ the complex extension of $L$, it is clear from $\S 3$ that $\lambda_{0}$ is an eigenvalue of $L$ if and only if $\lambda_{0}$ is a real eigenvalue of $M$. We define the generalized multiplicity of an eigenvalue $\lambda_{0} \neq 0$ of $L$ as the dimension $\mu\left(\lambda_{0}\right)$ of the space

$$
\begin{equation*}
E_{\lambda_{0}}=\left\{x \in E \mid\left(\lambda_{0} I-L\right)^{n} x=\theta\right\} \tag{4.5}
\end{equation*}
$$

for some integer $n \geq 1 . \mu\left(\lambda_{0}\right)$ is finite, for it is clear that $\mu\left(\lambda_{0}\right) \leq \nu\left(\lambda_{0}\right)$, the generalized multiplicity of $\lambda_{0}$ as an eigenvalue of $M$ (see subsection 2.17; actually $\mu\left(\lambda_{0}\right)=\nu\left(\lambda_{0}\right)$ as we will see in subsection 4.11).
4.3. Definition. Let $l$ and $L$ be as in subsection 4.2. We define the index $j(l)$ to be 1 if $L$ has no eigenvalues $>1$. If $L$ has eigenvalues $>1$, we denote them by $\lambda_{1}>\lambda_{2}>\cdots>\lambda_{s}$ and define the index $j(l)$ by

$$
\begin{equation*}
j(l)=(-1)^{\sum_{\rho=1}^{\theta} \mu_{\rho}} \tag{4.6}
\end{equation*}
$$

where $\mu_{\rho}=\mu\left(\lambda_{\rho}\right), \rho=1,2, \ldots, s$. Note that this definition agrees with (1.5), which was shown to agree in the finite-dimensional case with (1.2). Note also that $\lambda=1$ is not an eigenvalue of $L$ since $l$ is not singular.
4.4. Definition. A linear map $l$ of the real Banach space $E$ into itself is said to be type $I^{-}$if it can be constructed as follows: let $E^{1}$ be a one-dimensional subspace of $E$ and let $E^{2}$ be a complementary subspace such that every $x \in E$ can be written in the form (cf. Lemma 4 of Chapter 1)

$$
\begin{equation*}
x=x_{1}+x_{2}, \quad x_{1} \in E^{1}, x_{2} \in E^{2} \tag{4.7}
\end{equation*}
$$

Then $l$ maps $x=x_{1}+x_{2}$ into $-x_{1}+x_{2}$ (cf. $\S 8$ of the introduction). Maps of type $I^{-}$in a complex Banach space $Z$ are defined correspondingly.

Note that $l(x)=x-2 x_{1}$, i.e., $l(x)$ is of the form (4.3) with $L(x)=2 x_{1}$. Since $x_{1}$ lies in the one-dimensional space $E^{1}, L(x)$ is completely continuous. Thus $l$ is an L.-S. map. Since obviously $l$ maps $E$ onto $E$, the map $l$ is not singular. Thus $j(l)$ is defined for a map $l$ of type $I^{-}$.
4.5. Lemma.

$$
j(l)= \begin{cases}+1 & \text { if } l=I, \text { the identity map }  \tag{4.8}\\ -1 & \text { if } l \text { is of type } I^{-}\end{cases}
$$

Proof. If $l=I$ then the $L(x)$ in (4.3) equals $\theta$ for all $x$. Therefore (4.4) for $x \neq \theta$ is satisfied only with $\lambda_{0}=0$. Thus $L$ has no eigenvalues $\geq 1$ which by Definition 4.3 proves the first part of assertion (4.8). Suppose now $l$ is of the $I^{-}$type. Then with the notation used in (4.7), $l(x)=-x_{1}+x_{2}=x-2 x_{1}$, i.e., $L(x)=2 x_{1}$ by (4.3) and the eigenvalue equation (4.4) reads $2 x_{1}=\lambda_{0} x=$ $\lambda_{0}\left(x_{1}+x_{2}\right)$. This implies $x_{2}=\theta$ and $\lambda_{0}=2$. Thus 2 is the only eigenvalue of $L$, and

$$
\begin{equation*}
j(l)=(-1)^{\mu(2)} \tag{4.9}
\end{equation*}
$$

Here $\mu(2)$ is the dimension of the space given by (4.5) with $\lambda_{0}=2$ and $L(x)=$ $2 x_{1}$. Now for $i=1,2$ let $P_{i}$ be the projection $x \rightarrow x_{i}$ (cf. (4.7)). Then $\lambda_{0} I-L=$ $2\left(P_{1}+P_{2}\right)-2 P_{1}=2 P_{2}$. Thus, by (4.5), $E_{2}$ consists of those $x$ for which $P_{2}^{n} x=0$ for some integer $n \geq 1$. But for these integers $P_{2}^{n}=P_{2}$ since $P_{2}$ is a projection, and we see that $E_{2}=\left\{x \in E \mid P_{2}(x)=\theta\right\}=E^{1}$. Thus $\mu(2)=\operatorname{dim} E^{1}=1$. This together with (4.9) proves the second part of the assertion (4.8).

Since L.-S. homotopy (see subsection 1.2 of Chapter 2) is transitive, it follows from Lemma 4.5 that the linear homotopy theorem (stated in subsection 1.3 of Chapter 2) is equivalent to the following one.
4.6. ThEOREM. Let $l_{0}$ and $l_{1}$ be two linear nonsingular L.-S. maps $E \rightarrow E$. Then $l_{0}$ and $l_{1}$ are linearly L.-S. homotopic (see Definition 1.2 in Chapter 2) if and only if

$$
\begin{equation*}
j\left(l_{0}\right)=j\left(l_{1}\right) . \tag{4.10}
\end{equation*}
$$

Now for questions of continuity it is preferable to deal with the complexification $Z$ of $E$ and the complex extension $M$ of the operator $L$ given by (4.3). For if $M$ changes continuously, a real eigenvalue of $M$ may become complex. But in this case the eigenvalue of $L$ as an operator on $E$ "disappears" since the concept of a complex eigenvalue makes no sense in the real space $E$. We therefore state the following theorem whose proof will be given in $\S 5$.
4.7. Theorem. Let $l_{0}$ and $l_{1}$ be as in Theorem 4.6. Let $Z$ be the complexification of $E$, and for $i=0,1$ let $M_{i}$ be the complex extension of $L_{i}=I-l_{i}$ on $E$, and let $m_{i}(z)=z-M_{i}(z)$. It is asserted that the equality

$$
\begin{equation*}
j\left(m_{0}\right)=j\left(m_{1}\right) \tag{4.11}
\end{equation*}
$$

(see Definition 4.1) holds if and only if there exists an L.-S. linear homotopy $m(x, t)=z-M(z, t), 0 \leq t \leq 1$, between $m_{0}(z)=m(z, 0)$ and $m_{1}(z)=m(z, 1)$ which satisfies for $z \in Z$ and $t \in[0,1]$

$$
\begin{equation*}
M(\bar{z}, t)=\overline{M(z, t)} . \tag{4.12}
\end{equation*}
$$

The remainder of $\S 4$ is devoted to showing
4.8. Theorem. Theorems 4.6 and 4.7 are equivalent.

The proof of this theorem requires some lemmas.
4.9. Lemma. Let $E$ and $Z$ be as in Theorems 4.7 and 4.8. Let $L$ be a continuous linear map $E \rightarrow E$, and let $M$ be its complex extension. Then (i) the spectrum $\sigma(M)$ of $M$ is symmetric with respect to the real axis of the complex $\lambda$-plane, and (ii) for $\lambda \in \rho(M)$

$$
\begin{equation*}
R_{\bar{\lambda}}(M)(z)=\overline{R_{\lambda}(M)(\bar{z})} \tag{4.13}
\end{equation*}
$$

(see Definition 2.1).
Proof. (i) $\sigma(M)$ and the resolvent set $\rho(M)$ are disjoint subsets of the $\lambda$-plane whose union is the whole $\lambda$-plane. Therefore assertion (i) is equivalent to the assertion that $\rho(M)$ is symmetric to the real $\lambda$-axis. Let $\lambda_{0} \in \rho(M)$.

We have to prove

$$
\begin{equation*}
\bar{\lambda}_{0} \in \rho(M) \tag{4.14}
\end{equation*}
$$

Now by the definition of $R_{\lambda}(M)$

$$
\begin{equation*}
\left(\lambda_{0} I-M\right) R_{\lambda_{0}}(M)(z)=z \tag{4.15}
\end{equation*}
$$

for all $z \in Z$. But by (3.10) for $u \in Z$

$$
\overline{\left(\lambda_{0} I-M\right)(u)}=\left(\bar{\lambda}_{0} I-M\right)(\bar{u})
$$

Applying this relation with $u=R_{\lambda_{0}}(M)(z)$ we see from (4.15) that

$$
\left(\bar{\lambda}_{0} I-M\right) \overline{R_{\lambda_{0}}(M)(z)}=\bar{z}
$$

Replacing $z$ here by $\bar{z}$ we see that

$$
\left(\bar{\lambda}_{0} I-M\right) \overline{R_{\lambda_{0}}(M)(\bar{z})}=z
$$

for all $z \in Z$. This shows that

$$
\begin{equation*}
S(z)=\overline{R_{\lambda_{0}}(M)(\bar{z})} \tag{4.16}
\end{equation*}
$$

is a right inverse of $\left(\bar{\lambda}_{0} I-M\right)$. But interchanging the factors in (4.15) we conclude that $S(z)$ is also a left inverse of $\bar{\lambda}_{0} I-M$. Thus $S$ is an inverse of $\bar{\lambda}_{0} I-M$. Since $S$ is bounded, the assertion (4.14) follows by definition of $\rho(M)$.

Proof of (ii). By definition, $R_{\bar{\lambda}_{0}}(M)$ is the inverse of $\bar{\lambda}_{0} I-M$. Therefore by (4.16)

$$
R_{\bar{\lambda}_{0}}(M)(z)=S(z)=\overline{R_{\lambda_{0}}(M)(\bar{z})}
$$

This proves assertion (ii).
4.10. Lemma. Let $E, Z, L, M$ be as in Lemma 4.9. Let $\lambda_{0}$ be an isolated point of the spectrum $\sigma(M)$ of $M$ such that by Lemma $4.9 \bar{\lambda}_{0}$ is also an isolated point of $\sigma(M)$. Let $c_{0}$ be a counterclockwise-oriented circle with center $\lambda_{0}$ and radius $r_{0}$ where $r_{0}$ is such that the closure of the circular disk $B\left(\lambda_{0}, r_{0}\right)$ contains no point of $\sigma(M)$ other than $\lambda_{0}$. Let $c_{0}^{\prime}$ be the counterclockwise-oriented circle with center $\bar{\lambda}_{0}$ and radius $r_{0}$. Let $P_{\lambda_{0}}$ and $P_{\bar{\lambda}_{0}}$ be the linear operators defined by

$$
\begin{equation*}
P_{\lambda_{0}}=\frac{1}{2 \pi i} \int_{c_{0}} R_{\mu}(M) d \mu, \quad P_{\bar{\lambda}_{0}}=\frac{1}{2 \pi i} \int_{c_{0}^{\prime}} R_{\mu}(M) d \mu \tag{4.17}
\end{equation*}
$$

Then the ranges $Z_{\lambda_{0}}=P_{\lambda_{0}} Z$ and $Z_{\bar{\lambda}_{0}}=P_{\bar{\lambda}_{0}} Z$ are conjugate to each other, i.e., if $z$ varies over $Z_{\lambda_{0}}$, then $\bar{z}$ varies over $Z_{\bar{\lambda}_{0}}$.

PROOF. $\lambda_{0}$ and $\bar{\lambda}_{0}$ are spectral sets (see subsections 2.1 and 2.3). Therefore, by subsection 2.11, $P_{\lambda_{0}}$ and $P_{\bar{\lambda}_{0}}$ are projections. We will show first our lemma follows from the relation

$$
\begin{equation*}
P_{\bar{\lambda}_{0}}(\bar{z})=\overline{P_{\lambda_{0}}(z)} \quad \text { for all } z \in Z \tag{4.18}
\end{equation*}
$$

Indeed, since $P_{\lambda_{0}}$ is a projection, $z \in Z_{\lambda_{0}}=P_{\lambda_{0}} Z$ if and only if $z=P_{\lambda_{0}}(z)$ or $\bar{z}=\overline{P_{\lambda_{0}}(z)}$, or, by (4.18), $\bar{z}=P_{\bar{\lambda}_{0}}(\bar{z})$. But since $P_{\bar{\lambda}_{0}}$ is a projection, the last equality is a necesary and sufficient condition for $\bar{z}$ to be an element of $P_{\bar{\lambda}_{0}} Z=Z_{\bar{\lambda}_{0}}$.

Thus for the proof of our lemma it remains to verify (4.18). For this purpose we set $\mu=\bar{\lambda}_{0}+r_{0} e^{i \phi}$ and $\lambda=\bar{\mu}=\lambda_{0}+r_{0} e^{-i \phi}$ such that $d \mu=i r_{0} e^{i \phi} d \phi$ and $d \lambda=-i r_{0} e^{-i \phi} d \phi$. We then see from (4.17) and (4.13) that

$$
\begin{aligned}
P_{\bar{\lambda}_{0}}(\bar{z}) & =\frac{1}{2 \pi} \int_{0}^{2 \pi} R_{\mu}(M)(\bar{z}) r_{0} e^{i \phi} d \phi=\frac{1}{2 \pi} \int_{0}^{2 \pi} \overline{R_{\lambda}(M)(z)} r_{0} e^{i \phi} d \phi \\
& =\frac{1}{2 \pi} \int_{0}^{2 \pi} R_{\lambda}(M)(z) r_{o} e^{-i \phi} d \phi=-\frac{1}{2 \pi i} \int_{0}^{2 \pi} R_{\lambda}(M)(z)\left(-i r_{0} e^{-i \phi}\right) d \phi \\
& =-\frac{1}{2 \pi i} \int_{-c_{0}} R_{\lambda}(M)(z) d \lambda=\overline{\frac{1}{2 \pi i} \int_{c_{0}} R_{\lambda}(M)(z) d \lambda}=\overline{P_{\lambda_{0}}(z)} .
\end{aligned}
$$

This proves (4.18), and thus our lemma.
4.11. Corollary to Lemma 4.10. (i) The spaces $Z_{\lambda_{0}}$ and $Z_{\bar{\lambda}_{0}}$ have the same dimension; thus $\nu\left(\lambda_{0}\right)=\nu\left(\bar{\lambda}_{0}\right)$ (cf. subsection 2.17).
(ii) If $\lambda_{0}$ is real, then $Z_{\lambda_{0}}$ is invariant under conjugation.
(iii) If $\lambda_{0}$ is real, then the generalized multiplicity $\nu\left(\lambda_{0}\right)$ of $\lambda_{0}$ as an eigenvalue of $M$ equals the generalized multiplicity $\mu\left(\lambda_{0}\right)$ of $\lambda_{0}$ as an eigenvalue of $L$ (cf. subsections 2.17 and 4.2).

Proof. Assertions (i) and (ii) are obvious consequences of Lemma 4.10. As to assertion (iii) we noted already in subsection 4.2 that $\mu\left(\lambda_{0}\right) \leq \nu\left(\lambda_{0}\right)$. It remains to prove that

$$
\begin{equation*}
\nu\left(\lambda_{0}\right) \leq \mu\left(\lambda_{0}\right) \tag{4.19}
\end{equation*}
$$

Now it follows from assertion (ii) of our corollary and Lemma 3.12 that $Z_{\lambda_{0}}=$ $P_{\lambda_{0}} Z$ has a base $b_{1}, b_{2}, \ldots, b_{r}$ consisting of elements of $E$. Then

$$
r=\nu\left(\lambda_{0}\right)
$$

since $r$ and $\nu\left(\lambda_{0}\right)$ both equal the dimension of $Z_{\lambda_{0}}$ (cf. subsection 2.17). Since each $b_{j} \in Z_{\lambda_{0}}$, we have (also by subsection 2.17)

$$
\begin{equation*}
\left(\lambda_{0} I-M\right)^{n} b_{j}=\theta, \quad j=1,2, \ldots, \nu\left(\lambda_{0}\right) \tag{4.20}
\end{equation*}
$$

for some $n$. But by definition (3.9) of the extension $M$ of $L$, we see that $M\left(b_{j}\right)=$ $L\left(b_{j}\right)$ since $b_{j} \in E$. Thus by (4.20) $\left(\lambda_{0} I-L\right)^{n} b_{j}=0, j=1,2, \ldots, \nu\left(\lambda_{0}\right)$. The
asserted inequality (4.19) now follows from the definition of $\mu\left(\lambda_{0}\right)$ given in 4.2 since $\lambda_{0}$ is real.
4.12. Lemma. We use the notation of Lemma 4.10. We assume now that the linear map $l(x)=x-L(x)$ is a nonsingular L.-S. map $E \rightarrow E$. Let $m(z)$ be its complex extension. Then

$$
\begin{equation*}
j(m)=j(l) \tag{4.21}
\end{equation*}
$$

PROOF. By Lemma 3.9 the nonsingularity of $l$ implies that $m(z)=z-M(z)$ is a nonsingular L.-S. map $Z \rightarrow Z$. Thus both members of (4.21) are defined. To prove their equality we note first: if $\lambda_{0}$ is a complex eigenvalue of $M$, then so is $\bar{\lambda}_{0}$ (see Lemma 4.9), and $\nu\left(\lambda_{0}\right)=\nu\left(\bar{\lambda}_{0}\right)$ by part (i) of Corollary 4.11. Thus $\nu\left(\lambda_{0}\right)+\nu\left(\bar{\lambda}_{0}\right)$ is an even number. Therefore in Definition 4.1 (see in particular (4.2)) we may omit those $\nu_{i}=\nu\left(\lambda_{i}\right)$ for which $\lambda_{i}$ is complex. But for a real eigenvalue $\lambda_{i}$ of $M$ we know from part (i) of the corollary 4.11 that $\nu\left(\lambda_{i}\right)=\mu\left(\lambda_{i}\right)$, and the definition of $j(l)$ in subsection 4.3 (see in particular (4.6)) shows that $j(m)=j(l)$.
4.13. We are now ready to prove Theorem 4.8. In the present section we show that Theorem 4.7 implies Theorem 4.6.
(a) Suppose (4.10) is satisfied. We have to show that $l_{0}$ and $l_{1}$ are linearly L.-S. homotopic. Let $m_{0}$ and $m_{1}$ be the complex extensions of $l_{0}$ and $l_{1}$ resp. Then, by Lemma 4.12, the equality (4.10) implies (4.11). Therefore, by Theorem 4.7 there exists a homotopy $m(z, t)=z-M(z, t)$ of the properties asserted in that theorem. But by Lemma 3.11 the relation (4.12) implies that for each $t \in[0,1]$ the map $M(z, t)$ is the complex extension of a bounded linear map $L(x, t): E \rightarrow E$ which by Lemma 3.9 is completely continuous and such that $l(x, t)=x-L(x, t)$ is nonsingular. This $l(x, t)$ gives the desired homotopy between $l_{0}$ and $l_{1}$.
(b) Assume now that $l_{0}$ and $l_{1}$ are linearly L.-S. homotopic. We have to prove (4.10). By assumption there exists a linear L.-S. homotopy $l(x, t)=x-L(x, t)$ connecting $l_{0}$ with $l_{1}$. For each $t \in[0,1]$ let $m(x, t)=z-M(z, t)$ be the complex extension of $l(x, t)$. Then $m(z, t)$ is an L.-S. homotopy connecting $m_{0}(z)=$ $m(z, 0)$ with $m_{1}=m(z, 1)$ which by Lemma 3.11 satisfies (4.12). The assertion (4.10) follows now from Theorem 4.7 in conjunction with Lemma 4.12.
4.14. In this section we show that Theorem 4.6 implies Theorem 4.7.
(a) For $i=0,1$ let $m_{i}(z)=z-M_{i}(z)$ be a linear nonsingular L.-S. map $Z \rightarrow Z$ which is the complex extension of the linear nonsingular L.-S. map $l_{i}(x)=x-L_{i}(x): E \rightarrow E$, and suppose that (4.11) holds. Now by Lemma 4.12 the relation (4.11) implies (4.10). Consequently by Theorem 4.6 there exists a linear L.-S. homotopy $l(x, t)=x-L(x, t)$ connecting $l_{0}=l(x, 0)$ with $l_{1}=l(x, 1)$. Then the complex extension $m(z, t)=z-M(z, t)$ of $l(x, t)$ is a linear L.-S. homotopy which connects $m_{0}$ with $m_{1}$ and which, by Lemma 3.11, satisfies (4.12).
(b) Again, for $i=0,1$, let $m_{i}(z)=z-M_{i}(z)$ be a nonsingular linear L.-S. map $Z \rightarrow Z$ which is the extension of a linear nonsingular L.-S. map $l_{i}(x)=$ $x-L_{i}(x): E \rightarrow E$. Suppose now there exists a linear L.-S. homotopy which
connects $m_{0}$ with $m_{1}$ and satisfies (4.12). From the latter relation and from Lemma 3.11 we conclude that there exists a linear L.-S. homotopy which connects $l_{0}$ and $l_{1}$ and that therefore by Theorem 4.6, $j\left(l_{0}\right)=j\left(l_{1}\right)$. The asserted relation (4.11) follows now from Lemma 4.12.

## §5. Proof of the linear homotopy theorem

5.1. The linear homotopy theorem (see subsection 1.3 of Chapter 2) is, as already stated in the last paragraph of subsection 4.5, equivalent to Theorem 4.6, which in turn, by Theorem 4.8, is equivalent to Theorem 4.7. It is thus sufficient to prove Theorem 4.7. We start by proving that if a homotopy $m_{t}(z)=m(x, t)$ of the properties stated in that theorem exists, then (4.11) is true. To do this we will show that $j\left(m_{t}\right)$ is constant for $t \in[0,1]$. Since $j\left(m_{t}\right)$ is integer-valued it will be sufficient to prove the "continuity" of that number or, more precisely, to prove the following theorem.
5.2. TheOrem. Let $Z$ be the complexification of the real Banach space $E$, and let $m_{0}(z)=z-M_{0}(z)$ be the complex extension of the nonsingular linear $L .-S$. map $l_{0}(x)=x-L_{0}(x)$ mapping $E$ into $E$. Then there exists a positive number $\varepsilon_{0}$ of the following property: if $m(z)=z-M(z)$ is the complex extension of a linear nonsingular L.-S. map $l: E \rightarrow E$ and if

$$
\begin{equation*}
\left\|m-m_{0}\right\|=\left\|M-M_{0}\right\|<\varepsilon<\varepsilon_{0} \tag{5.1}
\end{equation*}
$$

then

$$
\begin{equation*}
j(m)=j\left(m_{0}\right) \tag{5.2}
\end{equation*}
$$

For the proof we need the following well-known elementary lemma.
5.3. Lemma. Let $Z$ be an arbitrary complex Banach space, and let $n_{0}$ be a bounded linear map $Z \rightarrow Z$ which has a bounded everywhere-defined inverse $n_{0}^{-1}$. Then every linear bounded map $n: Z \rightarrow Z$ satisfying

$$
\begin{equation*}
\left\|n-n_{0}\right\|<\left\|n_{0}^{-1}\right\|^{-1} \tag{5.3}
\end{equation*}
$$

has a bounded everywhere-defined inverse. (As the following proof shows, the lemma is also valid in a real Banach space E.)

Proof. Suppose first that $n_{0}=I$. Then the assumption (5.3) reads

$$
\begin{equation*}
\|n-I\|<1 \tag{5.4}
\end{equation*}
$$

from which it easily follows that

$$
\begin{equation*}
n_{1}=\sum_{i=0}^{\infty}(I-n)^{i} \tag{5.5}
\end{equation*}
$$

converges (with $(I-n)^{0}=I$ ). Moreover, writing $n=I-(I-n)$, one sees that $n n_{1}=n_{1} n=I$, and thus $n_{1}=n^{-1}$.

In the general case we see from (5.3) that

$$
\begin{equation*}
\left\|I-n_{0}^{-1} n\right\|=\left\|n_{0}^{-1}\left(n_{0}-n\right)\right\| \leq\left\|n_{0}^{-1}\right\|\left\|n_{0}-n\right\|<1 . \tag{5.6}
\end{equation*}
$$

Thus (5.4) is satisfied if $n$ is replaced by $n_{0}^{-1} n$. Thus $n_{2}=\left(n_{0}^{-1} n\right)^{-1}$ exists and $\left(n_{2} n_{0}^{-1}\right) n=n_{2}\left(n_{0}^{-1} n\right)=I$. This shows that $n_{2} n_{0}^{-1}$ is a left inverse of $n$. Similarly, we see from the inequality $\left\|I-n n_{0}^{-1}\right\|<1$, which is proved the same way as (5.6), that $n$ has a left inverse. But the existence of a left and a right inverse proves the existence of a unique inverse.
5.4. Proof of Theorem 5.2. We distinguish two cases: (A) $M_{0}$ has no eigenvalues $\lambda$ with real part $\mathcal{R}(\lambda) \geq 1$; (B) $M_{0}$ does have such eigenvalues.

Proof in case (A). We define the subset $\Delta=\Delta\left(M_{0}\right)$ of the $\lambda$-plane by $\Delta=$ $\{\lambda \mid R(\lambda) \geq 1\}$. Then $\Delta$ is a closed subset of $\rho\left(M_{0}\right)$. Therefore by Lemma 2.5 there exists a positive number $\mu_{0}$ such that

$$
\begin{equation*}
\left\|R_{\lambda}\left(M_{0}\right)\right\|<\mu_{0} / 2 \quad \text { for } \lambda \in \Delta \tag{5.7}
\end{equation*}
$$

We now claim that if $M$ is a linear completely continuous map $Z \rightarrow Z$ satisfying

$$
\begin{equation*}
\left\|M-M_{0}\right\|<\frac{1}{2}\left(\mu_{0} / 2\right)^{-1} \tag{5.8}
\end{equation*}
$$

then

$$
\begin{equation*}
\Delta \subset \rho(M) \tag{5.9}
\end{equation*}
$$

Indeed if $n_{0}(z)=\lambda z-M_{0}(z)$ and $n(z)=\lambda z-M(z)$ with $\lambda \in \Delta$, then by (5.8), by (5.7), and by the definition of $R_{\lambda}$

$$
\begin{equation*}
\left\|n-n_{0}\right\|=\left\|M-M_{0}\right\|<\left(\mu_{0} / 2\right)^{-1}<\left\|R_{\lambda}\left(M_{0}\right)\right\|^{-1}=\left\|n_{0}^{-1}\right\|^{-1} \tag{5.10}
\end{equation*}
$$

But by Lemma 5.3 this inequality implies that $n^{-1}=R_{\lambda}(M)$ exists, as a bounded operator, i.e., that $\lambda \in \rho(M)$ as asserted.

It is now easy to see that (5.8) implies the asserted equality (5.2). Indeed $j\left(m_{0}\right)=1$ by Definition 4.1. But by definition of $\Delta$ we see from (5.9) that no eigenvalue $\lambda$ of $M$ has a real part $\geq 1$. Thus, again by Definition 4.1, $j(m)=1$.

Proof in case (B). $\sigma\left(M_{0}\right)$ is closed and bounded (see subsection 2.1). Moreover by Lemma $3.9 M_{0}$ is completely continuous and therefore (see subsection 2.16) zero is the only accumulation point of $\sigma(M)$. It follows that $\sigma\left(M_{0}\right)$ contains only a finite number of points $\lambda$ with $R(\lambda) \geq 1$. Moreover, if we denote them by $\lambda_{1}^{0}, \lambda_{2}^{0}, \ldots, \lambda_{s}^{0}$, then there exists a positive number $r$ of the following properties: if for $i=1,2, \ldots, s, d_{i}$ is the open circular disk with center $\lambda_{i}^{0}$ and radius $r$, then (a) $\lambda_{i}^{0}$ is the only eigenvalue of $M_{0}$ in the closed disk $\bar{d}_{i}$; (b) the $\bar{d}_{i}$ are disjoint; (c) for those $i$ for which $R\left(\lambda_{i}^{0}\right)>1$ the disk $\bar{d}_{i}$ is contained in the open half plane $R(\lambda)>1$; (d) for those $i$ for which $\lambda_{i}^{0}$ is not real the disk $\bar{d}_{i}$ does not intersect the real axis; (e) for those $\lambda \in \sigma\left(M_{0}\right)$ for which $\mathcal{R}(\lambda)<1$ we have $\mathcal{R}(\lambda)<1-r$; (f) $0<r<1$. In our present case (B) we define

$$
\begin{equation*}
\Delta=\Delta\left(M_{0}\right)=\{\lambda \mid R(\lambda) \geq 1-r\}-\bigcup_{j=1}^{s} d_{j} \tag{5.11}
\end{equation*}
$$

Then $\Delta$ is a closed subset of $\rho\left(M_{0}\right)$. As in case (A) we conclude from this the existence of a positive $\mu_{0}$ such that (5.7) holds and for completely continuous linear $M$ satisfying (5.8) the inclusion (5.9) holds.

To prove that (5.1) with small enough $\varepsilon_{0}$ implies the assertion (5.2), we choose our notation in such a way that $\lambda_{1}^{0}, \lambda_{2}^{0}, \ldots, \lambda_{r}^{0}$ are real and $\lambda_{r+1}^{0}, \lambda_{r+2}^{0}, \ldots, \lambda_{s}^{0}$ are complex, $0 \leq r \leq s\left(r=0\right.$ means that there are no real numbers among the $\lambda_{i}^{0}$, and $r=s$ that all of them are real; in any case $s \geq 1$ by assumption). Since $M_{0}$ is the complex extension of a completely continuous operator $L_{0}: E \rightarrow E$, we see from Lemma 4.9 and the corollary to Lemma 4.11 that if $\lambda$ is one of the complex eigenvalues $\lambda_{r+1}^{0}, \ldots, \lambda_{0}^{s}$ of $M_{0}$, then so is $\bar{\lambda}$ and $\nu^{0}(\lambda)=\nu^{0}(\bar{\lambda})$, where $\nu^{0}(\lambda)$ denotes the generalized multiplicity of $\lambda$ as an eigenvalue of $M_{0}$. Consequently by definition (4.2) of $j(m)$

$$
\begin{equation*}
j\left(m_{0}\right)=(-1)^{\sum_{\rho=1}^{r} \nu\left(\lambda_{\rho}^{0}\right)} \tag{5.12}
\end{equation*}
$$

where it is understood that the sum stands for zero if $r=0$. Suppose now that $M$ is the complex extension of a linear completely continuous operator $L$ : $E \rightarrow E$ and that $M$ satisfies (5.8). Then (5.9) is true and therefore by definition (5.11) of $\Delta$

$$
\sigma(M) \cap\{\lambda \mid R(\lambda) \geq 1-r\} \subset \bigcup_{j=1}^{s} d_{j}=\bigcup_{j=1}^{r} d_{j}+\bigcup_{j=r+1}^{s} d_{j} .
$$

Now by property (d) of the radius $r$ of the disks $d_{j}$, these disks do not intersect the real axis for $j=r+1, \ldots, s$. Thus the eigenvalues of $M$ contained in these disks are complex, and the argument used to establish (5.12) shows that

$$
j(m)=(-1)^{\sum \nu(\lambda)}
$$

where the sum is extended over those eigenvalues $\lambda$ of $M$ which lie in one of the $d_{j}$ for $j=1,2, \ldots, r$ and where $\nu(\lambda)$ denotes the generalized multiplicity of $\lambda$ as an eigenvalue of $M$. Thus if $\lambda_{1}^{j}, \lambda_{2}^{j}, \ldots, \lambda_{r_{j}}^{j}$ are those eigenvalues of $M$ which lie in $d_{j}$ for $j=1,2, \ldots, r$, then

$$
\begin{equation*}
j(m)=(-1)^{\sum_{j=1}^{r} \sum_{i=1}^{r_{j}} \nu\left(\lambda_{i}^{j}\right)} . \tag{5.13}
\end{equation*}
$$

Comparison of (5.12) with (5.13) shows that our assertion (5.2) will be proved once it is shown that

$$
\begin{equation*}
\nu\left(\lambda_{j}^{0}\right)=\sum_{i=1}^{r_{j}} \nu\left(\lambda_{i}^{j}\right), \quad j=1,2, \ldots, r \tag{5.14}
\end{equation*}
$$

We recall that here the multiplicity $\nu$ at the left refers to $M_{0}$ while the $\nu$ 's at the right refer to $M$.

Now by subsection 2.17 and (2.15)

$$
\begin{equation*}
\nu\left(\lambda_{j}^{0}\right)=\operatorname{dim} P_{j}^{0} Z, \quad j=1,2, \ldots, r, \tag{5.15}
\end{equation*}
$$

where

$$
\begin{equation*}
P_{j}^{0}=\frac{1}{2 \pi i} \int_{\partial d_{j}} R_{\mu}\left(M_{0}\right) d \mu \tag{5.16}
\end{equation*}
$$

with the circle $\partial d_{j}$ oriented in the counterclockwise sense. In the same way we see that

$$
\begin{equation*}
\nu\left(\lambda_{i}^{j}\right)=\operatorname{dim} P_{i}^{j}, \quad i=1,2, \ldots, r_{j} \tag{5.17}
\end{equation*}
$$

where

$$
\begin{equation*}
P_{i}^{j}=\frac{1}{2 \pi i} \int_{\partial d_{i}^{j}} R_{\mu}(M) d \mu \tag{5.18}
\end{equation*}
$$

and where $d_{i}^{j}$ is a circular disk with center $\lambda_{i}^{j}$ and a radius so small that the closures $\bar{d}_{i}^{j}$ are disjoint and lie in $d_{j}$.

The assertion (5.14) now reads

$$
\begin{equation*}
\operatorname{dim} P_{j}^{0} Z=\sum_{i=1}^{r_{j}} \operatorname{dim} P_{i}^{j} Z, \quad j=1,2, \ldots, r \tag{5.19}
\end{equation*}
$$

If we set

$$
\begin{equation*}
P^{j}=\frac{1}{2 \pi i} \int_{\partial d_{j}} R_{\mu}(M) d \mu, \quad j=1,2, \ldots, r \tag{5.20}
\end{equation*}
$$

we see from the "Cauchy theorem" (2.4) that

$$
\begin{equation*}
P^{j}(Z)=\sum_{i=1}^{r_{j}} P_{i}^{j} Z \tag{5.21}
\end{equation*}
$$

and from this equality together with Lemma 2.12 that $P^{j} Z$ is the direct sum of the spaces $P_{1}^{j} Z, P_{2}^{j} Z, \ldots, P_{r_{j}}^{j} Z$, and that therefore

$$
\begin{equation*}
\operatorname{dim} P^{j} Z=\sum_{i=1}^{r_{j}} \operatorname{dim} P_{i}^{j} Z \tag{5.22}
\end{equation*}
$$

Thus the assertion (5.19) may be written as

$$
\begin{equation*}
\operatorname{dim} P_{j}^{0} Z=\operatorname{dim} P^{j} Z, \quad j=1,2, \ldots, r \tag{5.23}
\end{equation*}
$$

We will prove that if $\mu_{0}$ is as in (5.7), if $\varepsilon_{1}=\left(\mu_{0} / 2\right)^{-1}$ (cf. (5.8)), if

$$
\begin{equation*}
\varepsilon_{2}=\mu_{0}^{-2} \min _{j=1,2, \ldots, r}\left\|P_{j}^{0}\right\|^{-1} \tag{5.24}
\end{equation*}
$$

and if

$$
\begin{equation*}
\varepsilon_{0}=\min \left(\varepsilon_{1} / 2, \varepsilon_{2}\right) \tag{5.25}
\end{equation*}
$$

then (5.1) implies (5.23) (and thus (5.2)). For the proof we need two lemmas; the first is related to Lemma 5.3 while the second is due to J. Schwartz [50, p. 424].
5.5. Lemma. Let $n_{0}(z)$ be as in Lemma 5.3, and let $n$ be a linear bounded map $Z \rightarrow Z$ satisfying

$$
\begin{equation*}
\left\|n-n_{0}\right\|<\left(2\left\|n_{0}^{-1}\right\|\right)^{-1} \tag{5.26}
\end{equation*}
$$

Then $n^{-1}$ exists and satisfies

$$
\begin{equation*}
\left\|n^{-1}-n_{0}^{-1}\right\|<2\left\|n_{0}^{-1}\right\|^{2}\left\|n-n_{0}\right\| \tag{5.27}
\end{equation*}
$$

5.6. Lemma. Let $Q_{0}$ and $Q_{1}$ be projections $Z \rightarrow Z$. Suppose that

$$
\begin{equation*}
\left\|Q_{1}-Q_{0}\right\| \leq \frac{1}{2}\left\|Q_{0}\right\|^{-1} \tag{5.28}
\end{equation*}
$$

Then the dimension $\delta_{0}$ of $Q_{0} Z$ equals the dimension $\delta_{1}$ of $Q_{1} Z$.
5.7. We postpone the proof of the two preceding lemmas and prove that (5.1) implies (5.23). Let $n_{0}(z)=\mu z-M_{0}(z)$ and $n(z)=\mu z-M(z)$. Then by (5.8) and (5.7)

$$
\begin{aligned}
\left\|n(z)-n_{0}(z)\right\| & =\left\|M(z)-M_{0}(z)\right\|<\frac{1}{2}\left(\mu_{0} / 2\right)^{-1} \\
& <\frac{1}{2}\left\|R_{\mu}\left(M_{0}\right)\right\|^{-1}=\frac{1}{2}\left\|n_{0}^{-1}\right\|^{-1}
\end{aligned}
$$

i.e., (5.26) is satisfied. Therefore by Lemma 5.5 the inequality (5.27) holds:

$$
\left\|R_{\mu}(M)-R_{\mu}\left(M_{0}\right)\right\|<2\left\|R_{\mu}\left(M_{0}\right)\right\|^{2}\left\|M-M_{0}\right\|
$$

Thus by (5.1), (5.25) and (5.24)

$$
\left\|R_{\mu}(M)-R_{\mu}\left(M_{0}\right)\right\|<2\left\|R_{\mu}\left(M_{0}\right)\right\|^{2} \mu_{0}^{-2} \min _{j=1,2, \ldots, r}\left\|P_{j}^{0}\right\|^{-1}
$$

and by (5.7)

$$
\begin{equation*}
\left\|R_{\mu}(M)-R_{\mu}\left(M_{0}\right)\right\|<\frac{1}{2} \min _{j=1,2, \ldots, r}\left\|P_{j}^{0}\right\|^{-1} \tag{5.29}
\end{equation*}
$$

But by (5.20) and (5.16)

$$
\begin{aligned}
\left\|P^{j}-P_{j}^{0}\right\| & =\frac{1}{2 \pi}\left\|\int_{\partial d_{j}} R_{\mu}(M)-R_{\mu}\left(M_{0}\right) d \mu\right\| \\
& \leq \frac{1}{2 \pi} \int_{0}^{2 \pi}\left\|R_{\mu}(M)-R_{\mu}\left(M_{0}\right)\right\| r\left|e^{i \phi}\right| d \phi
\end{aligned}
$$

Therefore by (5.29) (recalling that $0<r<1$ ) we see that $\left\|P^{j}-P_{j}^{0}\right\|<\frac{1}{2}\left\|P_{j}^{0}\right\|^{-1}$. But by Lemma 5.6 this inequality implies the assertion (5.23).

It remains to prove Lemmas 5.5 and 5.6.
5.8. PROOF OF LEMMA 5.5. That the inequality (5.26) implies the existence of $n^{-1}$ follows from Lemma 5.3. Moreover it follows easily from the proof for that lemma (see in particular (5.5) and (5.6)) that

$$
n^{-1} n_{0}=\left(n_{0}^{-1} n\right)^{-1}=\sum_{j=0}^{\infty}\left(I-n_{0}^{-1} n\right)^{j}
$$

or

$$
n^{-1}=\sum_{j=0}^{\infty}\left(I-n_{0}^{-1} n\right)^{j} n_{0}^{-1}
$$

Thus

$$
n^{-1}-n_{0}^{-1}=\sum_{j=1}^{\infty}\left(I-n_{0}^{-1} n\right)^{j} n_{0}^{-1}=\left(I-n_{0}^{-1} n\right) \sum_{k=0}^{\infty}\left(I-n_{0}^{-1} n\right)^{k} n_{0}^{-1}
$$

Therefore

$$
\begin{aligned}
\left\|n^{-1}-n_{0}^{-1}\right\| & \leq\left\|n_{0}^{-1}\right\| \cdot\left\|I-n_{0}^{-1} n\right\| \cdot \sum_{k=0}^{\infty}\left\|I-n_{0}^{-1} n\right\|^{k} \\
& =\frac{\left\|n_{0}^{-1}\right\| \cdot\left\|I-n_{0}^{-1} n\right\|}{1-\left\|I-n_{0}^{-1} n\right\|} \leq \frac{\left\|n_{0}^{-1}\right\|^{2}\left\|n_{0}-n\right\|}{1-\left\|n_{0}^{-1}\right\| \cdot\left\|n_{0}-n\right\|}
\end{aligned}
$$

This inequality together with the assumption (5.26) proves the assertion (5.27).
5.9. Proof of Lemma 5.6. We will establish the inequalities

$$
\left.\begin{array}{l}
\left\|Q_{0}-Q_{0} Q_{1}\right\|  \tag{5.30}\\
\left\|Q_{1}-Q_{1} Q_{0}\right\|
\end{array}\right\}<1
$$

and then show that they imply our lemma. Since $Q_{0}^{2}=Q_{0}$ we see that

$$
\left\|Q_{0}-Q_{0} Q_{1}\right\|=\left\|Q_{0}\left(Q_{0}-Q_{1}\right)\right\| \leq\left\|Q_{0}\right\| \cdot\left\|Q_{0}-Q_{1}\right\|
$$

By assumption (5.28) this inequality implies the first of the two inequalities (5.30). To prove the second one, we note that $\left\|Q_{0}\right\| \geq 1$ since $Q_{0}$ is the identity on its range $Q_{0} Z$. Using this and again the assumption (5.28), we see that

$$
\begin{aligned}
\left\|Q_{1}\right\| & =\left\|Q_{0}+Q_{1}-Q_{0}\right\| \leq\left\|Q_{0}\right\|+\left\|Q_{1}-Q_{0}\right\| \\
& \leq\left\|Q_{0}\right\|+1 /\left(2\left\|Q_{0}\right\|\right) \leq\left\|Q_{0}\right\|+\left\|Q_{0}\right\| / 2
\end{aligned}
$$

Thus

$$
\begin{equation*}
\left\|Q_{1}\right\|<2\left\|Q_{0}\right\| \tag{5.31}
\end{equation*}
$$

Since $Q_{1}^{2}=Q_{1}$ we see that

$$
\left\|Q_{1}-Q_{1} Q_{0}\right\|=\left\|Q_{1}\left(Q_{1}-Q_{0}\right)\right\| \leq\left\|Q_{1}\right\| \cdot\left\|Q_{1}-Q_{0}\right\| .
$$

But if follows from (5.31) and from (5.28) that here the right member is less than 1. This proves the second of the equalities (5.30).

To derive the lemma from (5.30), we set $Z_{0}=Q_{0} Z, Z_{1}=Q_{1} Z$, and denote by $\left(Q_{0} Q_{1}\right)_{0}$ the restriction of $Q_{0} Q_{1}$ to $Z_{0}$. Then $\left(Q_{0} Q_{1}\right)_{0}$ maps $Z_{0} \rightarrow Z_{0}$, while $Q_{0}$ restricted to $Z_{0}$ is the identity on $Z_{0}$. Therefore the first of the inequalities (5.30) implies by Lemma 5.3 that $\left(Q_{0} Q_{1}\right)_{0}$ has an inverse on $Z_{0}$. Thus $Z_{0} \subset$ range of $\left(Q_{0} Q_{1}\right)_{0} \subset$ range of $Q_{0} Q_{1} \subset$ range of $Q_{1}$, and $Z_{0} \subset Q_{1} Z=Z_{1}$. Therefore $\delta_{0}=\operatorname{dim} Z_{0} \leq \delta_{1}=\operatorname{dim} Z_{1}$. In a similar way one derives from the second of the inequalities (5.30) that $\delta_{1} \leq \delta_{0}$. Thus $\delta_{1}=\delta_{0}$ as asserted.
5.10. We finished the proof that the equality (4.11) is necessary for the existence of the homotopy described in Theorem 4.7. We now turn to the sufficiency proof. We recall that $\lambda=1$ is not an eigenvalue of the complex extension $M_{0}$ of $L_{0}$ since $m_{0}=I-M_{0}$ is not singular. We consider first the special case that $M_{0}$ has no real eigenvalues $>1$. In this case it follows from Lemma 4.12 and Definition 4.3 that $j\left(m_{0}\right)=j\left(l_{0}\right)=1=j(I)$. Therefore in the special case considered we have to prove
5.11. Lemma. Let $M_{0}$ satisfy the assumption of Theorem 4.7. In addition it is assumed that $M_{0}$ has no real eigenvalues $\lambda>1$. Then there exists a linear L.-S. homotopy which connects $m_{0}$ with $I$ and which satisfies (4.12).

Proof. For $t \in[0,1]$ we set $m(z, t)=z-(1-t) M_{0}(z)$. That $M(z, t)=$ $(1-t) M_{0}(z)$ satisfies (4.12) is clear from Lemma 3.11. It remains to show that $m_{t}(z)=m(z, t)$ is nonsingular for all $t \in[0,1]$. For $t=0$ and $t=1$
the nonsingularity is obvious since $m(z, 0)=m_{0}(z)$ and $m(z, 1)=I$. Suppose then $m_{t_{0}}(z)$ to be singular for some $t_{0}$ in the open interval $(0,1)$. Then $z_{0}-$ $\left(1-t_{0}\right) M_{0}\left(z_{0}\right)=z_{0}$ for some $z_{0} \neq \theta$, or $M\left(z_{0}\right)=\lambda_{0} z_{0}$ with $\lambda_{0}=\left(1-t_{0}\right)^{-1}>1$. This contradicts our assumption that $M_{0}$ has no real eigenvalues $>1$.
5.12. We suppose now that $M_{0}$ has real eigenvalues greater than 1 . We denote them by $\lambda_{1}>\lambda_{2}>\cdots>\lambda_{s}$. Let $\sigma_{0}=\sigma_{0}\left(M_{0}\right)$ be their union, and let $\sigma_{1}=\sigma_{1}\left(M_{0}\right)$ be the complement of $\sigma_{0}$ in $\sigma\left(M_{0}\right)$ :

$$
\begin{equation*}
\sigma\left(M_{0}\right)=\sigma_{0} \cup \sigma_{1} \tag{5.32}
\end{equation*}
$$

Let

$$
\begin{equation*}
P_{0}=P_{\sigma_{0}}, \quad P_{1}=P_{\sigma_{1}} \tag{5.33}
\end{equation*}
$$

be the projections defined in Lemma 2.11 (cf. also subsection 2.12), and let

$$
\begin{equation*}
Z_{0}=P_{0} Z, \quad Z_{1}=P_{1} Z \tag{5.34}
\end{equation*}
$$

Then as will be shown in subsection 5.13

$$
\begin{equation*}
Z=Z_{0}+Z_{1} \quad \text { (direct sum) } \tag{5.35}
\end{equation*}
$$

with $Z_{0}$ being finite-dimensional. Corresponding to the decomposition (5.34) we set

$$
\begin{equation*}
M_{0}^{0}=P_{0} M_{0}, \quad M_{0}^{1}=P_{1} M_{0} \tag{5.36}
\end{equation*}
$$

We will show that

$$
M_{0}=\left(\begin{array}{cc}
M_{0}^{0} & 0  \tag{5.37}\\
0 & M_{0}^{1}
\end{array}\right)
$$

(see subsection 5.15) and that $\sigma_{0}$ is the spectrum of $M_{0}^{0}$ and $\sigma_{1}$ the spectrum of $M_{0}^{1}$ (see subsection 5.16). Thus $M_{0}^{1}$ has no real eigenvalues $>1$, and Lemma 5.11 may be applied to $m_{0}^{\prime}=I_{1}-M_{0}^{1}$, where $I_{1}$ denotes the identity map on $Z_{1}$. Discussion of the "finite-dimensional" part $M_{0}^{0}$ of $M_{0}$ will begin in subsection 5.20 .
5.13. Proof of (5.35). Let $V_{0}$ and $V_{1}$ be open sets in the $\lambda$-plane containing $\sigma_{0}$ and $\sigma_{1}$ resp. with $\bar{V}_{0}$ and $\bar{V}_{1}$ disjoint. Let $\Gamma_{0}=\partial V_{0}$ and $\Gamma_{1}=\partial V_{1}$, where $\Gamma_{0}$ and $\Gamma_{1}$ are supposed to be rectifiable Jordan curves oriented in the counterclockwise sense. Then by subsections 2.11 and 2.12, by (5.33) and by Lemma 2.9

$$
\begin{equation*}
I=P_{0}+P_{1}=\frac{1}{2 \pi i} \int_{\Gamma} R_{\mu}\left(M_{0}\right) d \mu, \quad \Gamma_{0}+\Gamma_{1} \tag{5.38}
\end{equation*}
$$

This shows that every $z \in Z$ can be represented by

$$
\begin{equation*}
z=z_{0}+z_{1}, \quad z_{0} \in Z_{0}, z_{1} \in Z_{1} \tag{5.39}
\end{equation*}
$$

The uniqueness of this representation follows from (2.12).
To establish (5.37) we prove the following lemma.
5.14. LEMMA.

$$
\begin{align*}
& M_{0}^{0}=\frac{1}{2 \pi i} \int_{\Gamma_{0}} \mu R_{\mu}\left(M_{0}\right) d \mu=M_{0} P_{0}  \tag{5.40}\\
& M_{0}^{1}=\frac{1}{2 \pi i} \int_{\Gamma_{1}} \mu R_{\mu}\left(M_{0}\right) d \mu=M_{0} P_{1}, \tag{5.41}
\end{align*}
$$

where $M_{0}^{0}$ and $M_{0}^{1}$ are the operators defined in (5.36). Moreover, for $z \in Z$,

$$
\begin{align*}
M_{0}(z) & =M_{0}^{0}\left(z_{0}\right)+M_{0}^{1}(z)  \tag{5.42}\\
M_{0}^{0}\left(z_{1}\right) & =\theta, \quad M_{0}^{1}\left(z_{0}\right)=\theta \tag{5.43}
\end{align*}
$$

with $z_{0}, z_{1}$ as in (5.39).
Proof. Let

$$
f(\mu)= \begin{cases}1 & \text { for } \mu \in \bar{V}_{0}, \\ 0 & \text { for } \mu \in \bar{V}_{1},\end{cases}
$$

where $V_{0}$ and $V_{1}$ are as in subsection 5.13 . Then by (2.10) and (5.33)

$$
P_{0}=\frac{1}{2 \pi i} \int_{\Gamma_{0}} R_{\mu}\left(M_{0}\right) d \mu=\frac{1}{2 \pi i} \int_{\Gamma_{0}+\Gamma_{1}} f(\mu) R_{\mu}\left(M_{0}\right) d \mu
$$

and from (5.32) and Lemma 2.9 we see that

$$
M_{0}=\frac{1}{2 \pi i} \int_{\Gamma_{0}+\Gamma_{1}} \mu R_{\mu}\left(M_{0}\right) d \mu
$$

Therefore from (2.8) with $g(\mu)=\mu$ on $\bar{V}_{0} \cup \bar{V}_{1}$

$$
\begin{equation*}
P_{0} M_{0}=M_{0} P_{0}=\frac{1}{2 \pi i} \int_{\Gamma_{0} \cup \Gamma_{1}} f(\mu) \mu R_{\mu}\left(M_{0}\right) d \mu=\frac{1}{2 \pi i} \int_{\Gamma_{0}} \mu R_{\mu}\left(M_{0}\right) d \mu \tag{5.44}
\end{equation*}
$$

By definition (5.36) this equality proves (5.40). Assertion (5.41) is proved correspondingly.

To prove (5.42) we note that by (5.39) and (5.34)

$$
\begin{aligned}
M_{0}(z) & =M_{0}\left(z_{0}+z_{1}\right)=M_{0}\left(z_{0}\right)+M_{0}\left(z_{1}\right) \\
& =M_{0}\left(P_{0} z\right)+M_{0}\left(P_{1} z\right)=M_{0} P_{0}\left(z_{0}\right)+M_{0} P_{1}\left(z_{1}\right)
\end{aligned}
$$

By (5.40) and (5.41) this proves (5.42). Finally by (5.36), (5.44), and (2.12)

$$
M_{0}^{0}\left(z_{1}\right)=P_{0} M_{0}\left(z_{1}\right)=M_{0} P_{0}\left(z_{1}\right)=M_{0} P_{0} P_{1}(z)=0 .
$$

This proves the first of the relations (5.43). The second one follows similarly.
5.15. Proof of (5.37). Since $P_{0} M_{0}=M_{0} P_{0}$ by (5.44) and since $P_{1} M_{0}=$ $M_{0} P_{1}$ is established the same way,

$$
M_{0}(z)=\binom{P_{0} M_{0}(z)}{P_{1} M_{0}(z)}=\binom{M_{0} P_{0}(z)}{M_{0} P_{1}(z)} .
$$

Therefore by the direct decomposition (5.39)

$$
M_{0}(z)=\left(\begin{array}{ll}
M_{0} P_{0}\left(z_{0}\right) & M_{0} P_{0}\left(z_{1}\right) \\
M_{0} P_{1}\left(z_{0}\right) & M_{0} P_{1}\left(z_{1}\right)
\end{array}\right)
$$

This relation proves (5.37) as we see from (5.40), (5.41), and (5.43).
5.16. Proof that $\sigma_{0}$ is the spectrum of $M_{0}^{0}$ and $\sigma_{1}$ is the spectrum of $M_{0}^{1}$. It follows from (5.42) and (5.43) that, restricted to $Z_{0}, M_{0}=M_{0}^{0}$, and, restricted to $Z_{1}, M_{0}=M_{0}^{1}$. Our assertion now follows from subsection 2.13.

This finishes the proof of the assertions contained in subsection 5.12. We next prove
5.17. Lemma. (i) $Z_{0}$ and $Z_{1}$ are invariant under conjugation (see Definition 3.10); (ii) $Z_{0}$ is finite dimensional.

Proof. (i) Let $\lambda_{j}, j=1,2, \ldots, s$, be as in subsection 5.12. Let $d_{j}$ be a circular disk with center $\lambda_{j}$ of radius so small that the closures $\bar{d}_{j}$ are disjoint and lie in $V_{0}$ (cf. subsection 5.13). Let the boundary $\partial d_{j}$ of $d_{j}$ be oriented in the counterclockwise sense, and let

$$
\begin{equation*}
Q_{j}=\frac{1}{2 \pi i} \int_{\partial d_{j}} R_{\mu}\left(M_{0}\right) d \mu, \quad j=1,2, \ldots, s \tag{5.45}
\end{equation*}
$$

Since the $\lambda_{j}$ are the only eigenvalues of $M_{0}$ in $V_{0}$, it follows from the Cauchy theorem 2.7 and from 5.12 that

$$
\begin{equation*}
P_{0}=\sum_{j=1}^{s} Q_{j} \tag{5.46}
\end{equation*}
$$

and

$$
\begin{equation*}
Z_{0}=P_{0} Z=\sum_{j=1}^{s} Q_{j} Z \tag{5.47}
\end{equation*}
$$

Since each $\lambda_{j}$ is a spectral set for $M_{0}$, each $Q_{j}$ is a projection by Lemma 2.11, and by the Corollary 4.11 to Lemma 4.10 each of the spaces $Q_{j} Z$ is invariant under conjugation since $\lambda_{j}$ is real. Thus we see from (5.47) that $Z_{0}$ is invariant under conjugation. Since $Z$ also is invariant under conjugation, being the complexification of a real Banach space $E$, it now follows that $Z_{1}$ also is invariant under conjugation.
(ii) By Lemma $2.12 Q_{i} Q_{j}=0$ for $i \neq j$. It follows that the decomposition (5.47) is direct. Therefore

$$
\operatorname{dim} Z_{0}=\sum_{j=1}^{s} \operatorname{dim} Q_{j} Z_{0}
$$

But by 2.17 the dimension of $Q_{j} Z_{0}$ is the finite number $\nu\left(\lambda_{j}\right)$. Thus

$$
\begin{equation*}
\operatorname{dim} Z_{0}=\sum_{j=1}^{s} \nu\left(\lambda_{j}\right) \tag{5.48}
\end{equation*}
$$

5.18. Let $I_{0}$ and $I_{1}$ be the identity maps on $Z_{0}$ and $Z_{1}$ resp., and let, as always, $I$ denote the identity map on $Z$. Then

$$
m_{0}=I-M_{0}\left(\begin{array}{cc}
m_{0}^{0} & 0  \tag{5.49}\\
0 & m_{0}^{1}
\end{array}\right)
$$

where

$$
\begin{equation*}
m_{0}^{0}=I_{0}-M_{0}^{0}, \quad m_{0}^{1}=I_{1}-M_{0}^{1} \tag{5.50}
\end{equation*}
$$

This is an immediate consequence of (5.37).
5.19. LEMMA. There exists an L.-S. homotopy $m_{0}^{1}\left(z_{1}, t\right)=z_{1}-M_{0}^{1}\left(z_{1}, t\right)$, $z_{1} \in Z_{1}, t \in[0,1]$, connecting $m_{0}^{1}$ with $I_{1}$ and satisfying

$$
M_{0}^{1}(\bar{z}, t)=\overline{M_{0}^{1}(z, t)} .
$$

Proof. The lemma will be proved once it is shown that $M_{0}^{1}$ satisfies the assumption made in Theorem 5.11 for $M_{0}$. That is, we have to verify: (i) $Z_{1}$ is the complexification of a real Banach space $E_{1}$; (ii) $m_{0}^{1}=I_{1}-M_{0}^{1}$ and $I_{1}$ are the complex extensions of linear nonsingular L.-S. maps on $E_{1}$; (iii) $M_{1}^{0}$ has no real eigenvalues $>1$. Now (i) follows from the fact that $Z_{1}$ is invariant under conjugation (Lemma 5.17). The assertion of (ii), that $I_{1}$ is the complex extension of the identity on $E_{1}$, is obvious. To prove that $M_{0}^{1}$ is the complex extension of a linear continuous map $L_{0}^{1}$ on $E_{1}$, it is by Lemma 3.11 sufficient to prove the relation

$$
\begin{equation*}
M_{0}^{1}(\bar{z})=\overline{M_{0}^{1}(z)}, \quad z \in Z_{1} . \tag{5.51}
\end{equation*}
$$

But this relation follows from the fact that $M_{0}^{1}(z)=M_{0}(z)$ for $z \in Z_{1}$, by (5.41) and that $M_{0}(\bar{z})=\overline{M_{0}(z)}$ since $M_{0}$ is by assumption the extension of a linear continuous map on $E$. That $L_{0}^{1}$ is completely continuous follows from Lemma 3.9 since $M_{0}^{1}(z)=M_{0}(z)$ is completely continuous by assumption. Finally, that $l_{0}^{1}=I_{1}-L_{0}^{1}$ is nonsingular follows also from Lemma 3.9 since otherwise, by that lemma, $m_{0}^{1}$ and therefore $m_{0}$ would be singular. Finally that assertion (iii) is true was already noticed in the lines directly following (5.37).
5.20. Our next goal is to prove that $m_{0}=I-M_{0}$ is L.-S. homotopic (with a homotopy satisfying (4.12)) to either $I$ or to a map of type $I^{-}$(see Definition 4.4). It follows from subsection 5.19 and (5.49) that $m_{0}$ is L.-S. homotopic (with the condition (4.12)) to

$$
\tilde{m}_{0}=\left(\begin{array}{cc}
m_{0}^{0} & 0  \tag{5.52}\\
0 & I_{1}
\end{array}\right), \quad m_{0}^{0}=I_{0}-M_{0}^{0}
$$

It will therefore be sufficient to prove the corresponding assertion for the map $m_{0}^{0}$ which maps the finite-dimensional space $Z_{0}$ onto $Z_{0}$ (cf. Lemma 5.17). Our first step in this direction will be to prove
5.21. LEMMA. There exists a real base $b_{1}, b_{2}, \ldots, b_{N}$ of $Z_{0}$ such that

$$
\begin{equation*}
m_{0}^{0}\left(b_{j}\right)=-b_{j}, \quad j=1,2, \ldots, N \tag{5.53}
\end{equation*}
$$

where $N$ is the dimension of $Z_{0}$ (cf. (5.48)).
Proof. As shown in the proof of part (ii) of Lemma 5.17 the decomposition of $Z_{0}$ given by (5.47) is direct. Therefore a basis for $Z_{0}$ is obtained by putting
together the bases for each of the spaces $Q_{j} Z_{0}, j=1,2, \ldots, s$. But by subsection 2.17 and by assertion (iii) of Corollary 4.11

$$
\begin{equation*}
\operatorname{dim} Q_{j} Z_{0}=\nu\left(\lambda_{j}\right)=\mu\left(\lambda_{j}\right) \tag{5.54}
\end{equation*}
$$

Now let $\lambda_{0}$ be one of the $\lambda_{j}$, and let $Q_{0}$ be the corresponding $Q_{j} . Q_{0} Z_{0}$ is invariant under conjugation (see Corollary 4.11). It follows that $Q_{0} Z_{0}$ is the complexification of a real Banach space $E_{0}$ of the same dimension, and every base of $E_{0}$ is a real base of $Z_{0}$ (cf. Lemma 3.12). Therefore if $L_{0}$ is the linear map $E_{0} \rightarrow E_{0}$ of which the map $M_{0}^{0}$ (restricted to $Q_{0} Z_{0}$ ) is the complex extension, it will for the proof of our lemma be sufficient to show that there exists a base $b_{1}, \ldots, b_{N_{0}}$ of $E_{0}$ such that

$$
\begin{equation*}
l_{0}\left(b_{j}\right)=-b_{j}, \quad j=1,2, \ldots, N_{0} \tag{5.55}
\end{equation*}
$$

where

$$
\begin{equation*}
l_{0}(x)=x-L_{0}(x), \quad x \in E_{0} \tag{5.56}
\end{equation*}
$$

Let $n_{1}$ be the "index of nilpotency" of the map $Q_{0} Z_{0} \rightarrow Q_{0} Z_{0}$ given by

$$
\begin{equation*}
l_{\lambda_{0}}=\lambda_{0} I_{0}-L_{0} \tag{5.57}
\end{equation*}
$$

i.e., the smallest positive integer $n$ for which (4.5) holds for all $x \in Q_{0} Z_{0}$. Then by a well-known theorem of linear algebra (see, e.g., [26, p. 111]) there exist integers $r, n_{2}, n_{3}, \ldots, n_{r}$ and elements $\beta_{1}, \beta_{2}, \ldots, \beta_{r}$ such that $n_{1} \geq n_{2} \geq \cdots \geq$ $n_{r}$ and such that the elements

$$
\begin{equation*}
\beta_{1}, l_{\lambda_{0}} \beta_{1}, \ldots, l_{\lambda_{0}}^{n_{1}-1} \beta_{1}, \ldots, \beta_{\rho}, l_{\lambda_{0}} \beta_{\rho}, \ldots, l_{\lambda_{0}}^{n_{\rho}-1} \beta_{\rho}, \ldots, \beta_{r}, l_{\lambda_{0}} \beta_{r}, \ldots, l_{\lambda_{0}}^{n_{r}-1} \beta_{r} \tag{5.58}
\end{equation*}
$$

form a base for $E_{0}$, while

$$
\begin{equation*}
l_{\lambda_{0}}^{n_{1}} \beta_{1}=l_{\lambda_{0}}^{n_{2}} \beta_{2}=\cdots=l_{\lambda_{0}}^{n_{r}} \beta_{r}=\theta \tag{5.59}
\end{equation*}
$$

Thus $E_{0}$ is the direct sum of spaces $E_{0 \rho}$ with bases

$$
\begin{equation*}
\beta_{\rho}^{1}=\beta_{\rho}, \quad \beta_{\rho}^{2}=l_{\lambda_{0}} \beta^{\rho}, \ldots, \beta^{n_{\rho}}=l_{\lambda_{0}}^{n_{\rho}-1} \beta_{\rho} \tag{5.60}
\end{equation*}
$$

We then see from (5.58) that

$$
\begin{equation*}
l_{\lambda_{0}} \beta_{\rho}^{1}=\beta_{\rho}^{2}, \quad l_{\lambda_{0}} \beta_{\rho}^{2}=\beta_{\rho}^{3}, \ldots, l_{\lambda_{0}} \beta_{\rho}^{n_{\rho}-1}=\beta_{\rho}^{n_{\rho}} \tag{5.61}
\end{equation*}
$$

while by (5.59)

$$
\begin{equation*}
l_{\lambda_{0}}^{n_{\rho}} \beta_{\rho}^{1}=\theta \tag{5.62}
\end{equation*}
$$

But from (5.56) and (5.57)

$$
l_{0}=\left(1-\lambda_{0}\right) I_{0}+l_{\lambda_{0}}
$$

and therefore from (5.61), (5.62),

$$
\begin{gather*}
l_{0} \beta_{\rho}^{1}=\left(1-\lambda_{0}\right) \beta_{\rho}^{1}+\beta_{\rho}^{2}, \ldots, l_{0} \beta_{\rho}^{n_{\rho}-1}=\left(1-\lambda_{0}\right) \beta_{\rho}^{n_{\rho}-1}+\beta_{\rho}^{n_{\rho}} \\
l_{0} \beta_{\rho}^{n_{\rho}}=\left(1-\lambda_{0}\right) \beta_{\rho}^{n_{\rho}} . \tag{5.63}
\end{gather*}
$$

We now define for $x \in E_{0 \rho}$ and $t \in[0,1]$ a homotopy $l_{t}=l(x, t)$ by setting

$$
\begin{align*}
& l_{t} \beta_{\rho}^{1}=\left[\left(1-\lambda_{0}\right)(1-t)+(-1) t\right] \beta_{\rho}^{1}+(1-t) \beta_{\rho}^{2}, \\
& \vdots  \tag{5.64}\\
& l_{t} \beta_{\rho}^{n_{\rho}-1}=\left[\left(1-\lambda_{0}\right)(1-t)+(-1) t\right] \beta_{\rho}^{n_{\rho}-1}+(1-t) \beta_{\rho}^{n_{\rho}}, \\
& l_{t} \beta_{\rho}^{n_{\rho}}=\left[\left(1-\lambda_{0}\right)(1-t)+(-1) t\right] \beta_{\rho}^{n_{\rho}} .
\end{align*}
$$

Obviously (5.63) and (5.64) agree for $t=0$, while for $t=1$ by (5.64)

$$
\begin{equation*}
l_{1} \beta_{\rho}^{j}=-\beta_{\rho}^{j}, \quad j=1,2, \ldots, n_{\rho} \tag{5.65}
\end{equation*}
$$

Recalling that $\lambda_{0}>1$ we see that $\left[\left(1-\lambda_{0}\right)(1-t)+(-1) t\right]$ is the linear convex combination of the points -1 and $1-\lambda_{0}<0$, and therefore does not contain the point 0 . This shows that $l_{t}$ is not singular for all $t \in[0,1]$. Thus $l_{0}$ and $l_{1}$ are homotopic and (5.65) proves our lemma
5.22. We will now prove the assertion stated in subsection 5.20 as our "next goal." By subsection 5.20 and Lemma 5.21 this assertion obviously follows from the following lemma.
5.23. Lemma. Let $Z_{0}$ be a complex Banach space of finite dimension $N$ which is the complexification of a real Banach space $E_{0}$. Let $m_{0}^{0}$ be a nonsingular linear map $Z_{0} \rightarrow Z_{0}$. Suppose there exists a real base $b_{1}, b_{2}, \ldots, b_{N}$ in $Z_{0}$ for which (5.54) holds. Then $m_{0}^{0}$ is L.-S. homotopic if (4.12) holds to the identity I if $N$ is even and to a map of type $I^{-}$if $N$ is odd.

Proof. If $N$ is even, we set $N=2 p$ and define $m_{0}^{0}(x, t)$ for $t \in[0,1]$ and for $i=1,2, \ldots, p$ by

$$
\begin{gather*}
m_{0}^{0}\left(b_{2 i-1}, t\right)=b_{2 i-1} \cos (1-t) \pi-b_{2 i} \sin (1-t) \pi  \tag{5.66}\\
m_{0}^{0}\left(b_{2 i}, t\right)=b_{2 i-1} \sin (1-t) \pi+b_{2 i} \cos (1-t) \pi
\end{gather*}
$$

Then $m_{0}^{0}(z, 0)=m_{0}^{0}(z)$ by (5.54), while $m_{0}^{0}(z, 1)$ sends $b_{j}$ into $b_{j}$ for $j=$ $1,2, \ldots, N$, i.e., $m_{0}^{0}(z, 1)$ is the identity map. Moreover $m_{0}^{0}(z, t)$ is not singular, the determinant of (5.66) being 1.

If $N$ is odd, we change our notation by denoting the given base of $Z_{0}$ by $b_{0}, b_{1}, \ldots, b_{N}$. We then set $m_{0}^{0}\left(b_{0}, t\right)=-b_{0}$, while $m_{0}^{0}\left(b_{j}, t\right)$ for $j=1,2, \ldots, N=$ $2 p$ is given by (5.66). We thus obtain a nonsingular homotopy $m_{0}^{0}(z, t)$ connecting the map $m_{0}^{0}(z)$ given by $m_{0}^{0}\left(b_{j}\right)=-b_{j}$ for $j=0,1, \ldots, N$ with the map given by $m_{0}^{0}\left(b_{0}, 1\right)=-b_{0}, m_{0}^{0}\left(b_{j}, 1\right)=b_{j}, j=1,2, \ldots, N$. The latter map is by definition of the $I^{-}$type.
5.23. We thus proved the assertion of subsection 5.20 . It is clear that the identity map $I$ on $Z$ is not homotopic to a map $m$ of type $I^{-}$since for such $m$ the equality $j(m)=j(I)$ would hold (see 5.10 ). But by (4.21) and Lemma 4.5 this leads to the contradiction $+1=-1$. Thus to finish the proof of Theorem 4.7 it remains to prove that two maps of the $I^{-}$type are L.-S. homotopic.
5.24. LEmMA. Let $m_{0}(z)=z-M_{0}(z)$ and $M_{1}(z)=z-M_{1}(z)$ be linear L.-S. maps $Z \rightarrow Z$ of type $I^{-}$which are complex extensions of the real maps $l_{0}$ and $l_{1}$ resp. Then there exists an L.-S. homotopy $m(z, t)=z-M(z, t)$ connecting $m_{0}$ with $m_{1}$ and satisfying the condition (4.12).

Proof. Again let $E$ denote the real Banach space of which $Z$ is the complexification. $l_{0}$ and $l_{1}$ are then L.-S. linear maps $E \rightarrow E$ of type $I^{-}$. It is easily seen that it will be sufficient to show that $l_{0}$ and $l_{1}$ are L.-S. homotopic.

By Definition 4.4 there exists for $i=0,1$ a direct decomposition $E=E_{1}^{i}+E_{2}^{i}$, where $E_{1}^{i}$ is a one-dimensional subspace, and there exists a corresponding unique representation for $x \in E$

$$
x=x_{1}^{i}+x_{2}^{i}, \quad x_{1}^{i} \in E_{1}^{i}, x_{2}^{i} \in E_{2}^{i}
$$

such that

$$
\begin{equation*}
l_{0}(x)=-x_{1}^{0}+x_{2}^{0}=x-2 x_{1}^{0}, \quad l_{1}(x)=x-2 x_{1}^{1} \tag{5.67}
\end{equation*}
$$

Now let $e^{0}$ and $e^{1}$ be unit elements in $E_{1}^{0}$ and $E_{1}^{1}$ resp.:

$$
\begin{equation*}
\left\|e^{0}\right\|=\left\|e^{1}\right\|=1 \tag{5.68}
\end{equation*}
$$

Then by (5.67)

$$
\begin{array}{cl}
l_{0}\left(e^{0}\right)=-e^{0}, & l_{1}\left(e^{1}\right)=-e^{1} \\
l_{0}(x)=x-2 \alpha(x) e^{0}, & l_{1}(x)=x-2 \beta(x) e^{1} \tag{5.69}
\end{array}
$$

where $\alpha$ and $\beta$ are real-valued continuous functionals on $E$ satisfying

$$
\begin{equation*}
\alpha\left(e^{0}\right)=\beta\left(e^{1}\right)=1 \tag{5.70}
\end{equation*}
$$

For the proof that the maps (5.69) are L.-S. homotopic, we distinguish three cases:
(A) $e^{0}=e^{1}$;
(B) $e^{0}=-e^{1}$;
(C) $e^{0}$ and $e^{1}$ are linearly independent.

Case (A). We set

$$
\begin{equation*}
\gamma_{t}(x)=(1-t) \alpha(x)+t \beta(x), \quad t \in[0,1] \tag{5.71}
\end{equation*}
$$

and

$$
\begin{equation*}
l_{t}(x)=x-2 \gamma_{t}(x) e^{0} \tag{5.72}
\end{equation*}
$$

It is clear that for $t=0$ and $t=1$ this definition agrees with the one given by (5.69). It remains to verify that $l_{t}$ is nonsingular for each $t \in[0,1]$. If this were not true, then $l_{\bar{t}}(\bar{x})=\theta$ for some $\bar{t} \in[0,1]$ and $\bar{x} \in E$ different from $\theta$, i.e., by (5.72), $\bar{x}=a e^{0}$ for some real $a \neq 0$. Since $l_{\bar{t}}$ is linear, we see that $l_{\bar{t}}\left(e^{0}\right)=\theta$, and therefore from (5.72) that $e^{0}\left(1-2 \gamma_{\bar{t}}\left(e^{0}\right)\right)=0$. Thus

$$
\begin{equation*}
2 \gamma_{\bar{t}}\left(e^{0}\right)=1 \tag{5.72a}
\end{equation*}
$$

But since $e^{0}=e^{1}$ by assumption, we see from (5.70) and (5.71) that $\gamma_{\bar{t}}\left(e^{0}\right)=$ $(1-\bar{t}) \alpha\left(e^{0}\right)+\bar{t} \beta\left(e^{1}\right)=1$ in contradiction to (5.72a).

Case (B). This case reduces to case (A) since, by (5.69), $l_{1}(x)=x-2 \tilde{\beta}(x) e^{0}$ with $\tilde{\beta}(x)=\beta(-x)$, and since $\tilde{\beta}\left(e^{0}\right)=\beta\left(-e^{0}\right)=\beta\left(e^{1}\right)=1$.

Case (C). Let $E_{2}$ be the two-dimensional subspace of $E$ spanned by $e^{0}$ and $e^{1}$, and let $E_{3}$ be a complementary subspace to $E_{2}$ such that every $x \in E$ has the unique representation

$$
\begin{equation*}
x=x_{2}+x_{3}, \quad x_{2} \in E_{2}, x_{3} \in E_{3} . \tag{5.73}
\end{equation*}
$$

Now let $\tilde{r}_{t}$ be the rotation in $E_{2}$ determined by

$$
\begin{align*}
& \tilde{r}_{t}\left(e^{0}\right)=e^{0} \cos (t \pi / 2)+e^{1} \sin (t \pi / 2),  \tag{5.74}\\
& \tilde{r}_{t}\left(e^{1}\right)=-e^{0} \sin (t \pi / 2)+e^{1} \cos (t \pi / 2) .
\end{align*} \quad 0 \leq t \leq 1 .
$$

We note that

$$
\begin{equation*}
\tilde{r}_{1}\left(e^{0}\right)=e^{1} \tag{5.75}
\end{equation*}
$$

We extend $\tilde{r}_{t}$ to a map $r_{t}$ on $E$ by setting, for $x \in E$ and $t \in[0,1]$,

$$
\begin{equation*}
r_{t}(x)=x-P_{2}(x)+\tilde{r}_{t}\left(P_{2}(x)\right) \tag{5.76}
\end{equation*}
$$

where $P_{2}$ is the projection $x=x_{2}+x_{3} \rightarrow x_{2}$ (cf. (5.73)). This is indeed an extension of $\tilde{r}_{t}$ since $x=P_{2}(x)$ for $x \in E_{2}$. (5.76) is obviously an L.-S. map. To show that it is nonsingular we will prove that $r_{t}^{-1}$ exists: let $y=r_{t}(x)$. Since $P_{2}^{2}=P_{2}$ and since $r_{t}\left(P_{2}(x)\right) \in E_{2}$, we see from (5.76) that $P_{2}(y)=\tilde{r}_{t}\left(P_{2}(x)\right)$, and since $\tilde{r}_{t}$ obviously has an inverse, it follows that

$$
\begin{equation*}
P_{2}(x)=\tilde{r}_{t}^{-1}\left(P_{2}(y)\right) . \tag{5.77}
\end{equation*}
$$

On the other hand, if $P_{3}$ is the projection $x=x_{2}+x_{3} \rightarrow x_{3}$, we see from (5.76) that $P_{3}(x)=P_{3}(y)$. Thus by (5.77) $x=P_{2}(x)+P_{3}(x)=\tilde{r}_{t}^{-1}\left(P_{2}(y)\right)+P_{3}(y)$ which shows that the inverse $r_{t}^{-1}$ exists. We therefore may define for $x \in E$ and $t \in[0,1]$

$$
\begin{equation*}
l(x, t)=x-2 \alpha\left(r_{t}^{-1}(x)\right) r_{t}\left(e^{0}\right) \tag{5.78}
\end{equation*}
$$

with $\alpha$ as in (5.69). Then

$$
\begin{equation*}
l(x, 0)=l_{0}(x) \tag{5.79}
\end{equation*}
$$

since $r_{0}$ is the identity map. Moreover by (5.75)

$$
\begin{equation*}
l(x, 1)=x-2 \alpha\left(r_{1}^{-1}(x)\right) e^{1}=x-2 \alpha_{1}(x) e^{1} \tag{5.80}
\end{equation*}
$$

where

$$
\begin{equation*}
\alpha_{1}(x)=\alpha\left(r_{1}^{-1}(x)\right) \tag{5.81}
\end{equation*}
$$

To show that $l_{0}(x)$ and $l(x, 1)$ are L.-S. homotopic we have, because of (5.80), only to show that $l(x, t)$ is not singular. If this were not true, then by (5.78)

$$
\begin{equation*}
x=2 \alpha\left(r_{t}^{-1}(x)\right) r_{t}\left(e^{0}\right) \tag{5.82}
\end{equation*}
$$

for some $x \in E$ with $x \neq \theta$ and some $t \in[0,1]$. Thus $x=a r_{t}\left(e^{0}\right)$ with $a \neq$ 0 . Substitution in (5.82) and cancelling $a$ shows that $1=2 \alpha\left(r_{t}^{-1} r_{t}\left(e_{0}\right)\right)=$ $2 \alpha\left(e^{0}\right)$ which contradicts (5.70). Now we want to prove that $l_{0}$ and $l_{1}$ (see (5.69) and (5.70)) are L.-S. homotopic. Since we just proved that $l_{0}$ and $l(\cdot, 1)$ are
L.-S. homotopic, it remains to show that $l_{1}$ and $l(\cdot, 1)$ are L.-S. homotopic. But comparing (5.80) with the definition (5.69) of $l_{1}$ we see that we are in case (A) provided that

$$
\begin{equation*}
\alpha_{1}\left(e^{1}\right)=1 \tag{5.83}
\end{equation*}
$$

Now by (5.81), (5.75) and (5.70), $\alpha_{1}\left(e^{1}\right)=\alpha\left(r_{1}^{-1}\left(e^{1}\right)\right)=\alpha\left(\tilde{r}_{1}^{-1}\left(e^{1}\right)\right)=\alpha\left(e^{0}\right)=1$.
This finishes the proof of the linear homotopy theorem.

## §6. Two multiplication theorems for the indices

6.1. Theorem. Let $E$ be a real Banach space, and $l_{0}$ and $l_{1}$ be two nonsingular L.-S. maps $E \rightarrow E$. Then

$$
\begin{equation*}
j\left(l_{0} l_{1}\right)=j\left(l_{0}\right) \cdot j\left(l_{1}\right) \tag{6.1}
\end{equation*}
$$

Proof. Let $I$ be the identity map on $E$ and let $l^{-}$be a fixed L.-S. map of type $I^{-}$on $E$. Then by Lemma 4.5

$$
\begin{equation*}
j(I)=+1, \quad j\left(l^{-}\right)=-1 . \tag{6.2}
\end{equation*}
$$

By the linear homotopy theorem each of the maps $l_{0}$ and $l_{1}$ is L.-S. homotopic to exactly one of the maps $I$ and $l^{-}$. Denoting L.-S. homotopy by the symbol " $\sim$ " there are four cases:
(i) $l_{0} \sim I, \quad l_{1} \sim I$;
(ii) $l_{0} \sim I, \quad l_{1} \sim l^{-}$;
(iii) $l_{0} \sim l^{-}, \quad l_{1} \sim I$;
(iv) $l_{0} \sim l^{-}, \quad l_{1} \sim l^{-}$.

Then

$$
\begin{aligned}
& l_{0} l_{1} \sim I \cdot I=I \quad \text { in case }(\mathrm{i}), \\
& l_{0} l_{1} \sim I \cdot l^{-}=l^{-} \quad \text { in case (ii) }, \\
& l_{0} l_{1} \sim l^{-} I=l^{-} \quad \text { in case (iii) }, \\
& l_{0} l_{1} \sim l^{-} l^{-}=I \quad \text { in case (iv) } .
\end{aligned}
$$

By Theorem 4.6 and by (6.2) this implies

$$
\begin{align*}
& j\left(l_{0} l_{1}\right)=+1 \quad \text { in case (i), } \\
& j\left(l_{0} l_{1}\right)=-1 \quad \text { in case (ii), } \\
& j\left(l_{0} l_{1}\right)=-1 \quad \text { in case (iii), }  \tag{6.4}\\
& j\left(l_{0} l_{1}\right)=+1 \quad \text { in case (iv). }
\end{align*}
$$

On the other hand we see from (6.3), from Theorem 4.6 and from (6.2) that

$$
\begin{array}{lll}
j\left(l_{0}\right)=1, & j\left(l_{1}\right)=1 \quad \text { in case (i) }, \\
j\left(l_{0}\right)=1, & j\left(l_{1}\right)=-1 \quad \text { in case (ii) }  \tag{6.5}\\
j\left(l_{0}\right)=-1, & j\left(l_{1}\right)=+1 \quad \text { in case (iii) }, \\
j\left(l_{0}\right)=-1, & j\left(l_{1}\right)=-1 \quad \text { in case (iv). }
\end{array}
$$

Comparison of (6.4) with (6.5) makes the assertion (6.1) evident.
6.2. Theorem. Let $l$ be a nonsingular L.-S. map $E \rightarrow E$. Let $E=E_{1}+E_{2}$ such that every $h \in E$ has the unique representation

$$
\begin{equation*}
h=h_{1}+h_{2}, \quad h_{1} \in E_{1}, h_{2} \in E_{2} \tag{6.6}
\end{equation*}
$$

Let $l_{1}$ and $l_{2}$ be the restrictions of $l$ to $E_{1}$ and $E_{2}$ resp., and suppose that

$$
l(h)=\left(\begin{array}{cc}
l_{1}\left(h_{1}\right) & 0  \tag{6.7}\\
0 & l_{2}\left(h_{2}\right)
\end{array}\right)
$$

Then

$$
\begin{equation*}
j(l)=j\left(l_{1}\right) \cdot j\left(l_{2}\right) \tag{6.8}
\end{equation*}
$$

Proof. It is clear that $l_{1}$ and $l_{2}$ are linear nonsingular L.-S. maps $E_{1} \rightarrow E$ and $E \rightarrow E_{2}$ resp. Thus the right member of (6.8) is defined. Now let

$$
m_{1}=\left(\begin{array}{cc}
l_{1} & 0  \tag{6.9}\\
0 & I_{2}
\end{array}\right), \quad m_{2}=\left(\begin{array}{cc}
I_{1} & 0 \\
0 & l_{2}
\end{array}\right)
$$

where $I_{1}$ and $I_{2}$ are the identity maps on $E_{1}$ and $E_{2}$ resp. Then $l=m_{1} m_{2}$. Therefore, by Theorem 6.1, $j(l)=j\left(m_{1}\right) \cdot j\left(m_{2}\right)$. It remains to verify that

$$
\begin{equation*}
j\left(l_{1}\right)=j\left(m_{1}\right), \quad j\left(l_{2}\right)=j\left(m_{2}\right) \tag{6.10}
\end{equation*}
$$

By Definition 4.3 of the index $j$, it will for the proof of the first of these equalities be sufficient to show:
(i) a real number $\lambda \neq 0$ is an eigenvalue of $M_{1}=I-m_{1}$ if and only if $\lambda$ is an eigenvalue of $L_{1}=I_{1}-l_{1}$;
(ii) for such an eigenvalue $\lambda \neq 0$

$$
\begin{equation*}
\mu_{l_{1}}(\lambda)=\mu_{m_{1}}(\lambda) \tag{6.11}
\end{equation*}
$$

where $\mu_{l_{1}}(\lambda)$ and $\mu_{m_{1}}(\lambda)$ denote the generalized multiplicity of $\lambda$ as an eigenvalue of $L_{1}$ and $M_{1}$ resp. (see Definition 4.2).

Proof of (i). Let $\lambda \neq 0$ be an eigenvalue of $M_{1}=I-m_{1}$. Then there exists an $h \neq 0$ such that $\left(I-m_{1}\right) h=\lambda h$ or by (6.9) and (6.6) such that

$$
\begin{equation*}
L_{1} h_{1}=\lambda h_{1} \tag{6.12}
\end{equation*}
$$

and $0 h_{2}=\lambda h_{2}$. This implies $h_{2}=\theta$ since $\lambda \neq 0$. Thus $h=h_{1}$, and (6.12) shows that $\lambda$ is an eigenvalue of $L_{1}$. The converse follows even more easily.

Proof of (ii). Let $\lambda \neq 0$ be an eigenvalue of $L_{1}$ and therefore, by (i), an eigenvalue of $M_{1}=I-m_{1}$. By (6.9)

$$
\lambda I-M_{1}=\left(\begin{array}{cc}
\lambda I_{1}-L_{1} & 0 \\
0 & \lambda I_{2}
\end{array}\right)
$$

and for any positive integer $n$

$$
\left(\lambda I-M_{1}\right)^{n}=\left(\begin{array}{cc}
\left(\lambda I_{1}-L_{1}\right)^{n} & 0 \\
0 & \lambda^{n} I_{2}
\end{array}\right)
$$

From this, from (6.6) and from $\lambda \neq 0$ it follows that for $h \in E,\left(\lambda I-M_{1}\right)^{n} h=\theta$ if and only if $\left(\lambda I_{1}-L_{1}\right)^{n} h_{1}=\theta$ and $h_{2}=\theta$. This implies (6.11) by Definition 4.2, and therefore, by Definition 4.3, proves the first of our assertions (6.10). The second one is proved correspondingly.
6.3. Corollary to Theorem 6.2. Let $E, E_{1}, E_{2}, h, h_{1}, h_{2}$ be as in Theorem 6.2, and for $i=1,2$ let $\pi_{i}$ be the projection $h=h_{1}+h_{2} \rightarrow h_{i}$. Let $m$ be a nonsingular L.-S. map $E \rightarrow E$. Let $m_{1}$ denote the restriction of $\pi_{1} m$ to $E_{1}$, and let $m_{2}$ denote the restriction of $\pi_{2} m$ to $E_{2}$. We assume

$$
\begin{equation*}
\pi_{2} m\left(E_{1}\right)=\theta \tag{6.13}
\end{equation*}
$$

Then

$$
\begin{equation*}
j(m)=j\left(m_{1}\right) \cdot j\left(m_{2}\right) \tag{6.14}
\end{equation*}
$$

PROOF. Let $m_{12}\left(h_{2}\right)=\pi_{1} m\left(h_{2}\right)$. Then

$$
m(h)=\left(\begin{array}{cc}
m_{1}\left(h_{1}\right) & m_{12}\left(h_{2}\right)  \tag{6.15}\\
\theta & m_{2}\left(h_{2}\right)
\end{array}\right)
$$

We note first that $m_{2}$ is nonsingular, for otherwise $m_{2}\left(h_{2}\right)=0$ for some $h_{2} \neq \theta$. Then $m(h)=0$ for $h=h_{2}$ since then $h_{1}=\theta$ and $m_{12}\left(h_{2}\right)=\pi_{1} m\left(h_{2}\right)=\theta$. This contradicts the assumed nonsingularity of $m$. Similarly one sees that $m_{1}$ is not singular.

Now let for $t \in[0,1]$

$$
m^{t}(h)=\left(\begin{array}{cc}
m_{1}\left(h_{1}\right) & t m_{12}\left(h_{2}\right)  \tag{6.16}\\
\theta & m_{2}\left(h_{2}\right)
\end{array}\right)
$$

We prove that for each $t$ this map is not singular by showing that for arbitrary $k \in E$ the equation

$$
\begin{equation*}
m^{t}(h)=k \tag{6.17}
\end{equation*}
$$

has a solution. Now if $k=k_{1}+k_{2}$ with $k_{1} \in E_{i}$, then by (6.16) the equation (6.17) is equivalent to

$$
m_{1}\left(h_{1}\right)+t m_{12}\left(h_{2}\right)=k_{1}, \quad m\left(h_{2}\right)=k_{2}
$$

Since $m_{1}$ and $m_{2}$ are nonsingular, this system obviously has a solution. We thus see from (6.16) and (6.15) that $m=m^{1}$ is L.-S. homotopic to the map

$$
m^{0}=\left(\begin{array}{cc}
m_{1}\left(h_{1}\right) & \theta \\
\theta & m_{2}\left(h_{2}\right)
\end{array}\right)
$$

Consequently by Theorem 4.6, $j(m)=j\left(m^{0}\right)$. But $j\left(m^{0}\right)=j\left(m_{1}\right) \cdot j\left(m_{2}\right)$ by Theorem 6.2. This proves the assertion (6.14).

## Appendix B

## Proof of the Sard-Smale Theorem 4.4 of Chapter 2

1. The Theorem of Sard. Let $R^{n}$ and $R^{m}$ be real Euclidean spaces of finite dimension $n$ and $m$ resp. Let $U$ be an open subset of $R^{n}$, and let $f \in C^{r}(U): U \rightarrow R^{m}$, where $r$ is a positive integer satisfying

$$
\begin{equation*}
r>\max (0, n-m) . \tag{1}
\end{equation*}
$$

Then the set of critical values of $f$ (see Definition 17 in Chapter 1) is of Lebesgue measure 0 .

For a proof of and further literature on Sard's theorem we refer the reader to $[1, \S 15]$.
2. Corollary to Sard's Theorem. Let $E^{n}$ and $E^{m}$ be (real) Banach spaces of finite dimension $n$ and $m$ resp. Let $\Omega$ be a bounded open set in $E^{n}$ and $f \in C^{r}(\bar{\Omega})(c f . \S 1.16)$ where $r$ is a positive integer satisfying (1). Then the set of points in $E^{m}$ which are regular values for $f$ (see §1.17) is dense in $E^{m}$.

This is an obvious consequence of Theorem 1 since on the one hand a subset of $R^{m}$ of measure 0 contains no open set and therefore its complement contains a point in every open subset of $R^{m}$ and is thus dense in $R^{m}$, while on the other hand every finite-dimensional Banach space is linearly isomorphic to a Euclidean space of the same dimension (see, e.g., [18, p. 245]).
3. The proof given below for the Sard-Smale theorem is an adaptation to the simpler Banach space case of the proof given in $[\mathbf{1}, \S 16]$ for Smale's generalization of Sard's theorem to certain Banach manifolds. Using the preceding corollary we will prove a "local version" of the Sard-Smale theorem in Lemma 4 for maps of a very special form. The lemma in $\S 5$ will allow us to show that the local theorem holds also for the more general maps treated in §6. The latter result implies the Sard-Smale theorem as shown in §7.
4. Lemma. Let $\Pi_{1}, K$, and $K^{*}$ be Banach spaces, where $K$ is finite dimensional and where

$$
\begin{equation*}
\operatorname{dim} K^{*}=\operatorname{dim} K-p, \tag{2}
\end{equation*}
$$

with $p$ being a positive integer. Let $\Pi$ and $\Sigma$ be the Banach spaces defined as the direct products

$$
\begin{equation*}
\Pi=\Pi_{1} \dot{+} K, \quad \Sigma=\Pi_{1} \dot{+} K^{*} \tag{3}
\end{equation*}
$$

(cf. §1.3). Let $V$ be a bounded open subset of $\Pi$. Let $\psi$ be a $C^{s}$-map $V \rightarrow \Sigma$ where $s$ is an integer satisfying

$$
\begin{equation*}
s \geq p+1 \tag{4}
\end{equation*}
$$

We suppose moreover that the map $\psi: V \rightarrow W=\psi(V)$ is of the special form

$$
\begin{equation*}
\psi(\varsigma)=\varsigma_{1}+\chi(\varsigma) \tag{5}
\end{equation*}
$$

where

$$
\begin{equation*}
\varsigma_{=}=\varsigma_{1}+\zeta_{2}, \quad \zeta_{1} \in \Pi_{1}, \varsigma_{2} \in K \tag{6}
\end{equation*}
$$

and where the map $\chi: V \rightarrow K^{*}$ is completely continuous.
Now if $\varsigma^{0}$ is a point of $V$, then there exists an open neighborhood $V_{0}$ of $\varsigma^{0}$ such that

$$
\begin{equation*}
V_{0} \subset \bar{V}_{0} \subset V \tag{7}
\end{equation*}
$$

and such that the set $R\left(\bar{\psi}_{0}\right)$ of those points of $\Sigma$ which are regular values for the restriction $\bar{\psi}_{0}$ of $\psi$ to $V_{0}$ is dense and open.

Proof. Let $\zeta^{0}=\zeta_{1}^{0}+k^{0}, \zeta_{1}^{0} \in \Pi_{1}, k^{0} \in K$. Since $\zeta^{0}$ is an element of the open set $V$, there obviously exist positive numbers $\varepsilon_{1}$ and $\varepsilon_{2}$ such that the "rectangular" neighborhood $V_{0}=V_{1} \times V_{2}$ satisfies (7) if

$$
\begin{equation*}
V_{1}=\left\{\varsigma_{1} \in \Pi_{1} \mid\left\|\varsigma_{1}-\zeta_{1}^{0}\right\|<\varepsilon_{1}\right\}, \quad V_{2}=\left\{k \in K \mid\left\|k-k^{0}\right\|<\varepsilon_{2}\right\} . \tag{8}
\end{equation*}
$$

We will show first that the set $R\left(\psi_{0}\right)$ of regular values for the restriction $\psi_{0}$ of $\psi$ to $V_{0}$ is dense in $\Sigma$. For this purpose we consider an arbitrary point $\bar{\sigma}=\bar{\zeta}_{1}+\bar{k}^{*}$ in $\Sigma$ where $\bar{\zeta}_{1} \in \Pi_{1}$ and $\bar{k}^{*} \in K^{*}$. We have to show that every neighborhood $N$ of $\bar{\sigma}$ contains a point which is a regular value for $\psi_{0}$. Now this is obvious if the equation

$$
\begin{equation*}
\psi_{0}(\varsigma)=\bar{\sigma}=\bar{\zeta}_{1}+\bar{k}^{*} \tag{9}
\end{equation*}
$$

has no solution in $V_{0}$ since then, by definition, $\bar{\sigma}$ is a regular value of $\psi_{0}$. Suppose now (9) has a solution $\tilde{\varsigma}_{1}+\tilde{k}, \tilde{\varsigma}_{1} \in V_{1}, \tilde{k} \in V_{2}$. Then by (5) and (9), $\tilde{\varsigma}_{1}=\bar{\zeta}_{1}$ and

$$
\begin{equation*}
\chi\left(\bar{\varsigma}_{1}+\tilde{k}\right)=\bar{k}^{*} \tag{10}
\end{equation*}
$$

Now for $\bar{\zeta}_{1} \in V_{1}$ we define a map $\chi_{\bar{\zeta}_{1}}: V_{2} \subset K$ into $K^{*}$ by setting

$$
\begin{equation*}
\chi_{\bar{\varsigma}_{1}}(k)=\chi\left(\bar{\varsigma}_{1}+k\right) \tag{11}
\end{equation*}
$$

From (2) and (4) we see that this map satisfies the assumption of the Corollary 2 (with $E^{m}=K^{*}, E^{n}=K$ ). Therefore by that corollary every neighborhood $N_{2} \subset K^{*}$ of $\bar{k}^{*}$ contains a regular value $k^{*}$ for $\chi_{\overline{5}_{1}}$. We will show that for such $k^{*}$ the point $\sigma=\bar{\zeta}_{1}+k^{*} \in \Sigma$ is a regular value for $\psi_{0}$. This will obviously prove our assertion that every neighborhood $N$ of $\bar{\sigma}=\bar{\zeta}_{1}+\bar{k}^{*}$ contains a regular value for $\psi_{0}$. Now if the equation

$$
\begin{equation*}
\chi_{\bar{\varsigma}_{1}}(k)=k^{*} \tag{12}
\end{equation*}
$$

has no solution $k \subset V_{2}$, then the equation

$$
\begin{equation*}
\psi_{0}(\varsigma)=\bar{\zeta}_{1}+k^{*} \tag{13}
\end{equation*}
$$

has no solution $\varsigma \in V_{0}$ as is seen from (5) and (11). Thus in this case $\bar{\zeta}_{1}+k^{*}$ is a regular value for $\psi_{0}$.

Suppose now (12) has a solution $k \in V_{2}$. Then since $k^{*}$ is a regular value for $\chi_{\bar{\varsigma}_{1}}$, the differential $D \chi_{\bar{\varsigma}_{1}}(k ; \kappa), \kappa \in K$, is nonsingular. On the other hand, $\varsigma=$ $\bar{\zeta}_{1}+k$ is the corresponding solution of (13), and with $h=\lambda+\kappa, \lambda \in \Pi_{1}, \kappa \in K$, we see from (5) that

$$
D \psi_{0}(\varsigma ; h)=\left(\begin{array}{cc}
\lambda & \theta  \tag{14}\\
D \chi(\varsigma ; \lambda) & D \chi(\varsigma ; \kappa)
\end{array}\right)
$$

But $D \chi(\varsigma ; \kappa)=D \chi_{\bar{\varsigma}_{1}}(k ; \kappa)$ by (11). This shows that the differential (14) is not singular and therefore that $\bar{\zeta}_{1}+k^{*}$ is a regular value for $\psi_{0}$ for every solution $\zeta \in V_{1} \times V_{2}$ of (13).

This finishes the proof that $R\left(\psi_{0}\right)$ is dense in $\Sigma$. Now let $U_{0}$ be a rectangular neighborhood of $\zeta^{0}$ whose closure $\bar{U}_{0}$ is a subset of $V_{0}$. As follows directly from the definition of "regular value," the relation $\bar{U}_{0} \subset V_{0}$ implies that the set of regular values for the restriction $\bar{\psi}_{0}$ of $\psi$ to $\bar{U}_{0}$ contains the set $R\left(\psi_{0}\right)$ and is therefore dense in $\Sigma$. In other words, changing our notation, we can choose the rectangular neighborhood $V_{0}$ of $\varsigma^{0}$ satisfying (7) in such a way that $R\left(\bar{\psi}_{0}\right)$ is dense in $\Sigma$.

To prove our lemma it remains to show that $R\left(\bar{\psi}_{0}\right)$ is open or, what is the same, that its complement, the set $S\left(\bar{\psi}_{0}\right)$ of singular values for $\bar{\psi}_{0}$, is closed. Now it follows directly from the definition of "singular value" that $S\left(\bar{\psi}_{0}\right)=\bar{\psi}_{0}\left(S_{1}\right)$ if $S_{1}$ denotes the set of singular points of $\bar{\psi}_{0}$ in $\bar{V}_{0}$. But to prove that $S_{1}$ is closed, it will be sufficient to show that $S_{1}$ is closed as follows from the special form (5) of $\psi$ (cf. the proof of part (ii) of $\S 1.10$ ). The proof that $S_{1}$ is closed is quite similar to the proof of part (ii) of Lemma 19 in Chapter 1: let $\varsigma^{1}, \varsigma^{2}, \ldots$ be a convergent sequence of points in $S_{1}$. We have to prove that

$$
\bar{\zeta}=\lim _{i \rightarrow \infty} \varsigma^{i} \in S_{1}
$$

Since $\varsigma^{i} \in S_{1} \subset \bar{V}_{0}$ and since $\bar{V}_{0}$ is a closed set, we see that

$$
\bar{\zeta} \in \bar{V}_{0} \subset V
$$

Now if $\bar{\zeta} \notin S_{1}$ the differential $D \psi(\bar{\varsigma} ; h)$ would not be singular, i.e., $\bar{\zeta} \subset V$ would be a regular point for $\psi$. But then, by the continuity of the differential and by Lemma 5.3 of Appendix A, all points of some neighborhood of $\bar{\zeta}$ would be regular points for $\psi$. This contradicts the fact that every neighborhood of $\bar{\zeta}$ contains infinitely many of the singular points $\zeta^{i}$.
5. Lemma. Let $\Pi$ be a Banach space, let $E$ be a subspace of $\Pi$ of finite codimension $p$, let $Z$ be an open bounded subset of $\Pi$, and let $\phi=\phi(z)$ be an
L.-S. map $\bar{Z} \rightarrow E$. We suppose that $\phi \in C^{s}(\bar{z})$ for some integer $s \geq 1$. Let $z_{0}$ be a point of $Z$ and

$$
\begin{equation*}
l_{0}(\eta)=D \phi\left(z_{0} ; \eta\right) \tag{15}
\end{equation*}
$$

Moreover, let $K$ denote the kernel of $l_{0}$, i.e., $K=\left\{\eta \in \Pi \mid l_{0}(\eta)=\theta\right\}$. Then the following assertions (i) and (ii) are made:
(i) $K$ is finite dimensional, and there exists a subspace $\Pi_{1}$ of $\Pi$ such that

$$
\begin{equation*}
\Pi=\Pi_{1} \dot{+} K \tag{16}
\end{equation*}
$$

Moreover if $R \subset E$ is the range of $l_{0}$, then there exists a finite-dimensional subspace $K^{*}$ of $E$ with the properties:

$$
\begin{equation*}
\text { (a) } E=R \dot{+} K^{*} ; \quad \text { (b) } \quad \operatorname{dim} K^{*}=\operatorname{dim} K-p . \tag{17}
\end{equation*}
$$

Thus every $y \in E$ has the unique representation

$$
\begin{equation*}
y=r+k^{*}=\pi_{1}(y)+\pi_{2}(y), \quad r \in R, k^{*} \in K^{*} \tag{18}
\end{equation*}
$$

where $\pi_{1}$ and $\pi_{2}$ are the projections $y \rightarrow r$ and $y \rightarrow k^{*}$ resp.
(ii) Let $\Sigma$ be the direct sum of $\Pi_{1}$ and $K^{*}$ (see part (ii) of §1.3):

$$
\begin{equation*}
\Sigma=\Pi_{1} \dot{+} K^{*} \tag{19}
\end{equation*}
$$

such that each point $\sigma \in \Sigma$ has the unique representation

$$
\begin{equation*}
\sigma=\sigma_{1}+\sigma_{2}=\pi_{1}^{\Sigma}(\sigma)+\pi_{2}^{\Sigma}(\sigma), \quad \sigma_{1} \in \Pi_{1}, \sigma_{2} \in K^{*} \tag{20}
\end{equation*}
$$

where $\pi_{1}^{\Sigma}$ and $\pi_{2}^{\Sigma}$ are the projections $\sigma \rightarrow \sigma_{1}$ and $\sigma \rightarrow \sigma_{2}$ resp. It is asserted: there exists a linear L.-S. isomorphism $h$ of $E$ onto $\Sigma$ and an L.-S. $C^{s}$ isomorphism $h_{1}$ of some neighborhood $U \subset Z \subset \Pi$ of $z_{0}$ onto some neighborhood $V \subset \Pi$ of $h_{1}\left(z_{0}\right)$ such that the map

$$
\begin{equation*}
\psi(\varsigma)=h \phi\left(h_{1}^{-1}(\varsigma)\right), \quad \varsigma \in V, \tag{21}
\end{equation*}
$$

is of the form (5) where $\varsigma, \varsigma_{1}, \varsigma_{2}$ and $\chi$ are as described in the lines following (5).
Proof. (i) That $K$ is finite dimensional follows from Lemmas 12 and 18 in Chapter 1, and the existence of a $\Pi_{1}$ satisfying (16) follows from Lemma 4 in Chapter 1. Since $\phi$ is an L.-S. map whose domain is the subset $\bar{Z}$ of $\Pi$ and whose range lies in the subspace $E$ of $\Pi$, it may also be considered as an L.-S. map $\bar{Z} \rightarrow \Pi$. Therefore, by Lemma 12 of Chapter 1 there exists a finite-dimensional subspace $K_{1}^{*}$ of $\Pi$ such that

$$
\begin{equation*}
\Pi=R \dot{+} K_{1}^{*} \tag{22}
\end{equation*}
$$

Now $E \cap R=R$ since $R \subset E$. Therefore intersecting both members of (22) with $E$ we see that the finite-dimensional space $K^{*}=E \cap K_{1}^{*}$ satisfies (17a). Since $p$ is the codimension of $E$ in $\Pi$, comparison of (22) with (17a) shows that $\operatorname{dim} K^{*}=\operatorname{dim} K_{1}^{*}-p$. But by $\S 1.12, \operatorname{dim} K_{1}^{*}=\operatorname{dim} K$. This proves $(17 \mathrm{~b})$.

Proof of (ii). We note first that the restriction of $l_{0}$ to $\Pi_{1}$ maps $\Pi_{1}$ onto $R$. Indeed, if $r$ is a given element of $R$, there exists by definition of $R$ a $z \in \Pi$ such that $l_{0}(z)=r$. But by (16)

$$
\begin{equation*}
z=z_{1}+k, \quad z_{1} \in \Pi_{1}, k \in K \tag{23}
\end{equation*}
$$

Since $l_{0}(k)=\theta$ by definition of $K$, we see that $l_{0}\left(z_{1}\right)=r$. Thus $l_{0}$ as a map $\Pi_{1} \rightarrow R$ is a nonsingular L.-S. map $\Pi_{1}$ onto $R$. It follows from the argument given for the proof of part (v) of $\S 1.12$ that this map has an inverse $m: R \rightarrow \Pi_{1}$ which is a nonsingular L.-S. map. Now every $y \in E$ has the decomposition (18). If for $r=\pi_{1}(y), k^{*}=\pi_{2}(y)$ we define

$$
\begin{equation*}
h(y)=m(r)+k^{*} \tag{24}
\end{equation*}
$$

we obtain a map of $E$ onto the space $\Sigma$ defined by (19). Since $m(r)$ is linear and L.-S. and since $K^{*}$ is finite dimensional, it follows that $h$ is a linear nonsingular L.-S. map.

Using the decomposition (16) we set for $z=z_{1}+k \in Z, z_{1} \in \Pi_{1}, k \in K$,

$$
\begin{equation*}
h_{1}(z)=m \pi_{1} \phi(z)+k . \tag{25}
\end{equation*}
$$

Since $\phi(z) \in E$, it follows from (18) that $\pi_{1} \phi(z) \in R$. Therefore the first term of the right member of (25) is an element of $\Pi_{1}$, and $h_{1}(z) \in \Pi$ by (16). Moreover since $\phi$ and $m$ are L.-S. mappings and since $K$ is finite dimensional, we see from (25) that $h_{1}$ is an L.-S. map.

We prove next that $h_{1}$ maps some open neighborhood $U$ of $z_{0}$ onto some open neighborhood $V$ of $h_{1}\left(z_{0}\right)$ and that $h_{1}^{-1} \in C^{s}(V)$. For this purpose it will by the inversion theorem in $\S 1.21$ be sufficient to show that the differential of $h_{1}$ at $z_{0}$ is the identity map on $\Pi$, i.e., that

$$
\begin{equation*}
D h_{1}\left(z_{0} ; \eta\right)=\eta \tag{26}
\end{equation*}
$$

for all

$$
\begin{equation*}
\eta=\eta_{1}+\kappa \in \Pi, \quad \eta_{1} \in \Pi_{1}, \kappa \in K \tag{27}
\end{equation*}
$$

(cf. (16)). Now since $m$ and $\pi_{1}$ are linear, we see from (25) that

$$
\begin{equation*}
D h_{1}(z ; \eta)=m \pi_{1} D \phi(z ; \eta)+\kappa \tag{28}
\end{equation*}
$$

for all $z \in Z$. But for $z=z_{0}$ we see from (15) and from (27) that $D \phi\left(z_{0} ; \eta\right)=$ $l_{0}(\eta)=l_{0}\left(\eta_{1}\right)$ since $\kappa \in K$ and, therefore, $l_{0}(\kappa)=\theta$. Since $R$ is the range of $l_{0}$, we see from (18) that $\pi_{1} D \phi\left(z_{0}, \eta\right)=\pi_{1} l_{0}\left(\eta_{1}\right)=l_{0}\left(\eta_{1}\right)$. But $l_{0}\left(\eta_{1}\right)$ is the restriction of $l_{0}$ to $\Pi_{1}$ and, by definition, $m$ is its inverse. Therefore $m \pi_{1} D \phi\left(z_{0} ; \eta\right)=$ $m l_{0}\left(\eta_{1}\right)=\eta_{1}$. Thus by (28) $D h_{1}\left(z_{0} ; \eta\right)=\eta_{1}+\kappa$, which by (27) proves (26).

Then let $U$ and $V$ be sets of the properties stated above. We set

$$
\begin{equation*}
\zeta=h_{1}(z) \tag{29}
\end{equation*}
$$

If $z$ varies over $U$, then $\zeta$ varies over $V$. Since $\zeta \in \Pi$ we may write $\zeta=\zeta_{1}+\zeta_{2}$, $\varsigma_{1} \in \Pi_{1}, \varsigma_{2} \in K$ (cf. (16)). Then by (25) and (20)

$$
\begin{equation*}
\varsigma_{1}=m \pi_{1} \phi(z) \in \Pi_{1}, \quad \varsigma_{2}=k \in K \tag{30}
\end{equation*}
$$

Now if

$$
\begin{equation*}
\psi(\varsigma)=h \phi h_{1}^{-1}(\varsigma), \quad \varsigma \in V \tag{31}
\end{equation*}
$$

then with $z=h_{1}^{-1}(\varsigma)$ by (24), (18) and (30)

$$
\psi(\varsigma)=h \phi(z)=m \pi_{1} \phi(z)+\pi_{2} \phi(z)=\zeta_{1}+\pi_{2} \phi h_{1}^{-1}(\varsigma)
$$

With $\chi(\varsigma)=\pi_{2} \phi h_{1}^{-1}(\varsigma)$ this equality shows that $\psi$ is of the asserted form (5). Obviously $\chi(\varsigma)$ is bounded and lies in $K^{*}$. Since $K^{*}$ is finite dimensional, we see that $\chi$ is completely continuous. Thus assertion (ii) of Lemma 4 is proved.
6. Lemma. With the notation used in Lemma 5 we suppose that the assumptions of that lemma are satisfied. In addition we strengthen the assumption that $\phi \in C^{s}(\bar{Z})$ with $s \geq 1$ by requiring that

$$
\begin{equation*}
s \geq p+1 \tag{32}
\end{equation*}
$$

It is asserted that corresponding to each point $z_{0} \in Z$ there exists an open set $U_{0}$ which in addition to satisfying

$$
\begin{equation*}
z_{0} \in U_{0} \subset \bar{U}_{0} \subset Z \tag{33}
\end{equation*}
$$

has the following property: if $\bar{\phi}_{0}$ is the restriction of $\phi$ to $\bar{U}_{0}$, then the set $R\left(\bar{\phi}_{0}\right)$ of points in $E$ which are regular values for $\bar{\phi}_{0}$ is dense and open.

Proof. It follows from Lemma 5 together with the assumption (32) that all assumptions of Lemma 4 are satisfied. Consequently, the point

$$
\begin{equation*}
\varsigma^{0}=h_{1}\left(z_{0}\right) \in \Pi \tag{34}
\end{equation*}
$$

has an open neighborhood $V_{0}$ with the properties asserted in that lemma. We claim that then the set

$$
\begin{equation*}
U_{0}=h_{1}^{-1}\left(V_{0}\right) \tag{35}
\end{equation*}
$$

has the properties asserted in the present lemma. In the first place $U_{0}$ is open and $\bar{U}_{0}=h_{1}^{-1}\left(\bar{V}_{0}\right)$ as follows from Theorem 2.3 in Chapter 5. Next the following two assertions hold:
(a) for $y_{0} \in E$ the equation $\phi(z)=y_{0}$ has a solution $z \in \bar{U}_{0}$ if and only if for $\sigma_{0}=h\left(y_{0}\right) \in \Sigma$ the equation $\psi(\varsigma)=\sigma_{0}$ has a solution $\varsigma \in \bar{V}_{0}$;
(b) for $z_{0} \in \bar{U}_{0}$ the differential $D \phi\left(z_{0} ; \cdot\right)$ is not singular if and only if the differential $D \psi\left(h_{1}\left(z_{0}\right) ; \cdot\right)$ is not singular.
(a) follows directly from the relation (21) between $\psi$ and $\phi$ and the properties of $h$ and $h_{1}$.
(b) holds for the same reasons in conjunction with the chain rule. But (a) and (b) together imply that $R\left(\bar{\phi}_{0}\right)=h^{-1} R\left(\bar{\psi}_{0}\right)$.

Now $R\left(\bar{\psi}_{0}\right)$ is dense and open by Lemma 4. Therefore $h^{-1} R\left(\bar{\psi}_{0}\right)$ is dense and open since $h^{-1}$ is a linear one-to-one L.-S. map $\Sigma$ onto $E$.
7. Proof of the Sard-Smale theorem 4.4 of Chapter 2. Let $y_{0}$ be an arbitrary point of $E$. We have to prove that every neighborhood (in $E$ ) of $y_{0}$ contains a regular value for $\phi$. Now if the equation

$$
\begin{equation*}
\phi(z)=y_{0}, \quad y_{0} \in E-\phi(\partial Z) \tag{36}
\end{equation*}
$$

has no solution, then, by definition, $y_{0}$ is a regular value for $\phi$ and every neighborhood of $y_{0}$ contains a regular value, e.g., $y_{0}$.

Suppose now that the set $C_{0}$ of roots of (36) is not empty. $C_{0}$ is compact by Lemma 10 in Chapter 1 , and every $z_{0}$ has a neighborhood $U_{0}$ of the property
asserted in Lemma 6. As $z_{0}$ varies over $C_{0}$, these neighborhoods form an open covering of this compact set. Consequently there exists a finite subcovering, say, $U_{0}^{1}, U_{0}^{2}, \ldots, U_{0}^{q}$ such that (cf. (33)) for $i=1,2, \ldots, q$

$$
\begin{equation*}
U_{0}^{i} \subset \bar{U}_{0}^{i} \subset Z ; \quad C_{0} \subset U=\sum_{i=1}^{q} U_{0}^{i} \tag{37}
\end{equation*}
$$

If $\bar{\phi}_{i}$ is the restriction of $\phi$ to $\bar{U}_{0}^{i}$, then by Lemma 6 the set $R\left(\bar{\phi}_{i}\right)$ of regular values for $\bar{\phi}_{i}$ is dense and open in $E$. Consequently by Baire's density theorem (see e.g. [2; p. 108, Theorem $\left.V^{\prime}\right]$ ) the set

$$
\begin{equation*}
\Delta=\bigcap_{i=1}^{q} R\left(\bar{\phi}_{i}\right) \tag{38}
\end{equation*}
$$

is dense in $E$, and therefore in every neighborhood of $y_{0}$. We show next the existence of an open neighborhood $N_{0}$ of $y_{0}$ such that

$$
\begin{equation*}
\phi^{-1}\left(N_{0}\right) \subset U \tag{39}
\end{equation*}
$$

Indeed, otherwise, there would exist a sequence of positive numbers $\delta_{\nu}$ converging to 0 , of points $y_{\nu} \in B\left(y_{0}, \delta_{\nu}\right)$, and of points $x_{\nu}$ such that

$$
\begin{equation*}
\phi\left(x_{\nu}\right)=y_{\nu}, \quad x_{\nu} \in \bar{Z} \tag{40}
\end{equation*}
$$

but

$$
\begin{equation*}
x_{\nu} \notin U . \tag{41}
\end{equation*}
$$

Now since the $y_{\nu}$ converge to $y_{0}$ and since $\phi$ is an L.-S. map, it is easy to see that a subsequence of the $x_{\nu}$ converges to a point $x_{0}$ which by (40) satisfies $\phi\left(x_{0}\right)=y_{0}$, i.e.,

$$
\begin{equation*}
x_{0} \in \phi^{-1}\left(y_{0}\right)=C_{0} \tag{42}
\end{equation*}
$$

But by (41)

$$
\begin{equation*}
x_{0} \notin U \tag{43}
\end{equation*}
$$

since $U$ is open. But by (42) and (37), $x_{0} \in C_{0} \subset U$ which contradicts (43).
Then let $N_{0}$ be a neighborhood of $y_{0}$ satisfying (39), and let $\Delta_{1}=\Delta \cap N_{0}$. $\Delta_{1}$ is not empty since $\Delta$ is dense and $N_{0}$ open in $E$. We now show that every point $y_{1}$ of $\Delta_{1}$ is a regular value for $\phi$. If the equation

$$
\begin{equation*}
\phi(z)=y_{1} \tag{44}
\end{equation*}
$$

has no solutions, there is nothing to prove. But if it has a solution, then every solution lies in $U$ by (39), and therefore, by (37), in some of the sets $U_{0}^{i}$, say for $i=1,2, \ldots, q_{1}$ with $i \leq q_{1} \leq q$. Consider now for such $i$ a solution $z \in U_{0}^{i}$. Then $\phi(z)=\phi_{i}(z)=y_{1} \in \Delta=\bigcap_{j=1}^{q} R\left(\phi_{j}\right) \subset R\left(\phi_{i}\right)$, and $D \phi(z ; h)$ is not singular.

This finishes the proof that every point of $N_{0} \cap \Delta$ is a regular value for $\phi$, and therefore it finishes the proof of the Sard-Smale theorem since an $N_{0}$ satisfying (39) can be chosen as a subset of a given neighborhood of $y_{0}$.

## References

1. R. Abraham and J. Robbins, Transversal mappings and flows, Benjamin, New York-Amsterdam, 1967.
2. P. Alexandroff and H. Hopf, Topologie, Springer-Verlag, 1935.
3. H. Amann and S. A. Weiss, On the uniqueness of the topological degree, Math. Z. 130 (1973), 39-54.
4. S. Banach, Théorie des opérations linéaires, Warsaw, 1932.
5. R. Bonic and J. Frampton, Smooth functions on Banach manifolds, J. Math. Mech. 15 (1966), 877-898.
6. Yu. G. Borisovich, V. G. Zvyagin, and Y. I. Sapronov, Nonlinear Fredholm maps and the Leray-Schauder theory, Uspeki Mat. Nauk. 32 (1977), 3-54; English transl. in Russian Math. Surveys 32 (1977).
7. K. Borsuk, Drei Sätze über die n-dimensionale Sphäre, Fund. Math. 20 (1933), 177-190.
8. L. E. J. Brouwer, Beweis der Invarianz der Dimensionzahl, Math. Ann. 70 (1911), 161-165.
9.__, Über Abbildung von Mannigfaltigkeiten, Math. Ann. 71 (1912), 97-115.
9. F. Browder, Nonlinear operators and nonlinear equations of evolution in Banach spaces, Proc. Sympos. Pure Math., vol. 18, pt. 2, Amer. Math. Soc., Providence, R. I., 1976.
10. F. Browder and R. Nussbaum, The topological degree for noncompact nonlinear mappings in Banach spaces, Bull. Amer. Math. Soc. 74 (1968), 671-676.
11. F. Browder and W. Petryshyn, Approximation methods and the generalized topological degree for nonlinear mappings in Banach spaces, J. Funct. Anal. 3 (1969), 217-245.
12. R. Courant and D. Hilbert, Methods of mathematical physics. Vol. 1, Interscience, 1953.
13. J. Cronin, Fixed points and topological degree in nonlinear analysis, Math. Surveys, no. 11, Amer. Math. Soc., Providence, R. I., 1964.
14. K. Deimling, Nichtlineare Gleichungen und Abbildungsgrade, SpringerVerlag, 1974.
15. J. Dieudonné, Foundations of modern analysis, Academic Press, 1960.
16. J. Dugundji, An extension of Tietze's theorem, Pacific J. Math. 1 (1951), 353-367.
17. N. Dunford and J. T. Schwartz, Linear operators. Part I: General theory, Interscience, 1958.
18. G. Eisenack and C. Fenske, Fixpunkttheorie, Bibliographisches Institut, Mannheim, 1978.
19. K. Elworthy and A. Tromba, Differential structures and Fredholm maps on Banach manifolds, Proc. Sympos. Pure Math, vol. 15, Amer. Math. Soc., Providence, R. I., 1970.
20. C. Fenske, Analytische Theorie des Abbildungsgrades in Banach Rämmen, Math. Nachr. 48 (1971), 279-290.
21. K. Geba and A. Granas, Infinite dimensional cohomology theories, J. Math. Pures Appl. 52 (1973), 145-270.
22. A. Granas, Homotopy extension theorem in Banach spaces and some of its applications to the theory of nonlinear equations, Bull. Acad. Polon. Sci. Math. Astr. Phys. 7 (1959), 387-394. (Russian; English summary).
23. __, The theory of compact vector fields and some of its applications to topology of functional spaces. (I). Rozprawy Mat., Warsaw, 1962.
24. $\qquad$ , Points fixes pour les applications compactes: espaces de Lefschetz et la théorie de l'indice, Presses de l'Université de Montréal, Montréal, 1980.
25. P. Halmos, Finite dimensional vector spaces, Van Nostrand, Princeton, N. J., 1958.
26. H. Hopf, Topologie, lectures given at the University of Berlin, 1926/27, (unpublished). Notes by E. Pannwitz.
27. S. T. Hu, Theory of retracts, Wayne State Univ. Press, Detroit, Mich., 1965.
28. W. Hurewicz and H. Wallman, Dimension theory, Princeton Univ. Press, Princeton, N. J., 1941.
29. N. Jacobson, Lectures in abstract algebra. Vol. II: Linear algebra, Van Nostrand, Princeton, N. J., 1953.
30. M. Krasnoselski1̆, Topological methods in the theory of nonlinear integral equations, Gosudorstv. Izdat. Tehn.-Teor. Lit., Moscow, 1956; English transl., Macmillan, 1964.
31. L. Kronecker, Uber die Charakteristik von Funktionen Systemen, Monatsberichte der Kgl. Preussischen Akademie der Wissenschaften, 1878, 145-152.
32. J. Kurzweil, On approximation in real Banach spaces, Studia Math. 14 (1954), 213-231.
33. J. Leray, Topologie des espaces de Banach, C. R. Acad. Sci. Paris, 200 (1935), 1082-1084.
34. ___ La théorie des points fixes et ses applications en analyse, (Proc. Internat. Congr. of Math., Vol. 2, Cambridge, Mass., 1950) Amer. Math. Soc., Providence, R. I., 1952, pp. 202-208.
35. J. Leray and J. Schauder, Topologie et équations fonctionnelles, Ann. École Norm. 51 (1934), 45-78.
36. N. G. Lloyd, Degree theory, Cambridge Univ. Press, New York, 1978.
37. J. Mawhin, Equivalence theorems for nonlinear operator equations and coincidence theory for some mappings in locally convex topological vector spaces, J. Differential Equations 12 (1972), 610-636.
38. J. W. Milnor, Topology from the differentiable viewpoint, The University Press of Virginia, Charlottesville, Va., 1965.
39. H. Minkowski, Geometrie der Zahlen, Teubner, Leipzig und Berlin, 1910.
40. N. Moulis, Approximations de fonctions différentiable sur certain espaces de Banach, Thesis, Université de Paris, 1970.
41. M. Nagumo, A theory of degree of mapping based on infinitesimal analysis, Amer. J. Math 73 (1951), 485-496.
42. ___, Degree of mapping in convex linear topological spaces, Amer. J. Math. 73 (1951), 497-511.
43. L. Nirenberg, Topics in nonlinear functional analysis, Lecture Notes, 1973-74, Courant Inst. of Math. Sciences, New York University, New York, 1974. Notes by R. A. Artin.
44. E. H. Rothe, Über Abbildungsklassen von Kugeln des Hilbertschen Raumes, Compositio Math. 4 (1937), 294-307.
45. ___ Zur Theorie der topologischen Ordnung und der Vektorfelder in Banachschen Räumen, Compositio Math. 5 (1937), 177-197.
46. __ The theory of topological order in some linear topological spaces, Iowa State College J. Sci. 13 (1939), 373-390.
47. ___, Mapping degree in Banach spaces and spectral theory, Math. Z. 63 (1955), 195-218.
48. A. Sard, The measure of critical values of differentiable maps, Bull. Amer. Math. Soc. 48 (1942), 883-890.
49. J. T. Schwartz, Perturbations of spectral operators and applications. I: Bounded perturbations, Pacific J. Math. 4 (1954), 415-458.
50. $\qquad$ , Nonlinear functional analysis, Lecture Notes, Courant Inst. of Math. Sciences, New York University, New York, 1963-1964.
51. H. W. Siegberg, Brouwer degree; history and numerical computation (Sympos. Fixed point algorithms and complementarity problems, Univ. of Southampton, Southampton, 1979), North-Holland, 1980, pp. 389-411.
52. S. Smale, An infinite dimensional version of Sard's theorem, Amer. J. Math. 87 (1965), 861-866.
53. E. Zeidler, Vorlesungen uber nichtlineare Funktional Analysis. I: Fixpunktsätze, Teubner, Leipzig, 1976.

## Index

Affine map, 115
Antipodal points, 138
Ball, 13
Banach space, 13
Barycenter, 125
Barycentric subdivision, 125
Base of a vector space generated by a set, 24
Bilinear map, 18
Borsuk
theorem on odd maps, 192
Antipodensatz, 198
Borsuk-Ulam theorem, 198
Cauchy theorem (for operator-valued integrands), 201
Central projection, 146, 147
Chain rule for differentials, 18
Codimension, 14
Co-kernel, 17
Compact set, 15
Complementary space, 13
Complete continuity, 15, 203
Complex extension of a linear "real" operator, 205
Complexification of a real Banach space, 204
Component of an open set, 23
Conjugation, 206
Connected set, 21

Convex
hull, 22
set, 21
Coordinate transformation, 100, 101
$C^{r}$-map, 19
Critical
point, 19
value, 19
Differential, 18
Direct
sum, 13
summand, 13
Dugundji extension theorem, 22
Eigenelement, 204
Eigenvalue:
definition, 203
generalized multiplicity of, 204
multiplicity of, 204
Extension theorem for odd
maps, 185
Face of a simplex, 115
Finite layer map, 28
$\mathcal{F}(M)$, definition of, 201
Fredholm alternative, 17
General position of a point set, 113
Generating set for a vector space, 24
Great circle, great sphere, 138

Half-space, 117
Homotopy extension theorem, 176
Hopf extension theorem, 172
Hopf-Krasnoselskiĭ homotopy
theorem, 173
Hull of planes in $E^{n}, 113$
$I^{-}$type map, 209
Independence of points in $E^{n}, 112$
Index
of a linear nonsingular L.-S. map on a complex Banach space, 208 on a real Banach space, 27, 209
of an isolated root for mappings between spheres, 143
Intersection interpretation of degree and winding number, 84, 153
Invariance of domain, theorem of, 98
Inversion theorem, 20
Jordan-Leray theorem, 104
Kernel of a linear map, 17
Krasnoselskiî's lemma, 19
Layer map, 28
Leray-Schauder's
first lemma, 64
second lemma, 66
Leray-Schauder (L.-S.) maps, 15
generalized (g.L.-S.), 69
Linear
homotopy theorem, 26, 210
L.-S. homotopy, 26

Mazur lemma, 22
Multiplication theorem for indices, 228

Nonsingular linear map, 17
Order with respect to a point, 147
Orientation
of $E^{n}, 116$
of a simplex, 115, 116
of a sphere, 139

Poincaré-Bohl theorem, 76, 78
Polygon, simple polygon, 21
Precompact, 22
Product of Banach spaces, 13
Projection, 14, 202
Ray, 21
Real elements of a complexified Banach space, 204
Regular
linear map, 17
point, 19
value, 19
Resolvent set, 200
Rouché theorem, 78
Sard theorem, 231
Sard-Smale theorem, 40
Segment, 21
Simplex, 114
Simplicial
approximation, 126
map, 123, 153
set, 123
Singular
linear map, 17
point, 19
value, 19
(2-) Smooth Banach space, 63
Spectral set, 200
Spectrum, 200
Sphere, 138
Star, 125
domain, 79
Stereographic projection, 139
Subspace, 13
Sum theorem, 28
Tangent space, tangent plane, 139
Transversal, 22
Vector space generated by a set, 24
Winding number, 77

