

INTRODUCTION

A considerable effort has been devoted during the last years to the study of nonlinear diffusion equations or reaction-diffusion systems. A typical example is the following nonlinear parabolic problem

$$(1.1) \quad \begin{aligned} u_t &= -\Delta u + f(u), \quad \text{in } \Omega \times (0,T), \\ u_{|_{\partial\Omega}} &= 0, \quad \text{in } \Omega \times (0,T). \end{aligned}$$

where Ω is a smooth and compactly bounded domain in \mathbb{R}^N , $u|_{\partial\Omega} = 0$ are real numbers and f and g are sufficiently smooth functions. Of course, the above system should be complemented with suitable boundary and initial conditions. These systems, where the interaction between diffusion (measured by $-\Delta$) by the reaction (concentrations $u, g \geq 0$) and the reaction term f and g give rise to an extremely rich phenomenology, arise in many applications, giving qualitative models for a number of physical and biological phenomena. We can mention, among many others, chemical reactions (e.g., autocatalysis), combustion theory, population dynamics (predator-prey interactions, competition, etc.), nerve conduction (Kuramoto-Sivashinsky system), morphogenesis (cellular-vacuum-inhibitor interactions), superconductivity, etc. Cf. the book by Amann [1981] for more information and references (of, also, the lecture notes by Pata 1991).

One of the most fundamental problems posed by reaction-diffusion systems is the asymptotic behavior of solutions, i.e., to investigate what happens with solutions when the time t goes to infinity. (Clearly, this implies the existence and uniqueness of solutions.) From this point of view it is particularly interesting to study the existence and (large) multiplicity of solutions of the stationary problem associated to (1.1), namely the elliptic system

$$(1.2) \quad \begin{aligned} -u_{xx} + f(u,v) &= 0 & x \in \Omega \\ u_x(0) = u_x(1) &= 0 & \end{aligned}$$

together with the corresponding boundary condition. Indeed, in many interesting cases, the solutions of the evolution problem converge to one of the solutions (so-called steady state) of (1.2). Therefore, the study of the stability of stationary solutions is one of the most important problems in the theory. This is not the only possibility, and the reader may find a great variety of other possible situations (traveling waves, pulsating patterns, long-term periodic solutions, etc.) in a paper by Hale [55]. (However, a great deal of work has been devoted to the stability of solutions of systems as (1.2). It is well-known that oscillations of stability when parameter values are slightly varied with bifurcation phenomena, i.e., with changes in the structure of the solution set (1.2), when the parameters approach a critical value. We do not intend to trace this question here, but the reader can find some references in the references.)

The aim in these notes is to give two basic methods of solutions.

Functional analysis can be applied to the study of existence and multiplicity of steady states (1.2). We do not pretend to give an exhaustive survey of the different mathematical tools, and a similar consideration can be made concerning the variety of problems considered. On the contrary, our intention is to exhibit how some mathematical tools (the sub and superfunctions, degree theory, critical point theory, etc.) work in a few examples which seem interesting and useful for further research. However, this overview of qualitative methods

(compactness methods, topological methods, variational methods) is far from complete to mention only one important however, localization methods (in e.g., the Aubin-Lions-DStampfli method [34], [78], [181], [182], [183] (see also [184])) we must acknowledge. In other words,

it is very difficult than the methods in these notes have other interesting applications: for example, compactness arguments involving sub and superfunctions (cf., e.g., [182], [183], [48]), topological degree [52], local linearizations [48] and the fixed point index [181] have applications to the study of the stability of steady-state solutions, and kin and superfunctions [34] and degree theory [12] have been used in the study of travelling waves.

In particular attention is paid to the case of a single equation. Of course, it is more reasonable to deal with this case before to treat PDE systems, but there is also another reason for this: by the a developing technique, this section can be reduced to a single equation with a modest perturbation, and some of the proofs are very similar to the case of a single equation. Cf. Remark for more details.

Very often we are only interested in positive solutions when dealing with problems arising in physical applications. This concept will appear in several ways along these notes. In particular we will use positive arguments concerning the critical point in the corresponding function spaces and some related topics (the Mountain-Pass theorem, Nehari manifold) in the line of a series of positive functions, replacing the topological degree by the Land (just like, cf. the excellent survey by Adams [10] and [14]) for a very simple application of this kind of arguments.

The first chapter contains an exposition of the method of sub and superfunctions and some applications. Because 1.1 treats the case of a nonlinear elliptic equations, and the method is applied to Sect. 1.2 to the nonlinear eigenvalue problem

$$(1.3) \quad \begin{aligned} -u_{xx} + f(u) &= h & x \in \Omega \\ u_x(0) &= u_x(1) & \end{aligned}$$

where h is a real parameter and f is, however, rapidly oscillating, see (1.2).

In stark contrast of positive solutions, a unique non-trivial solution, say the 1/2 derivative in modulus of the metric to express well-known, interesting asymptotic and wave applications.

Chapter 2 is devoted to topological methods (especially the degree) Schröder degree and norm of the applications. Section 2.1 gives a brief history without proofs of the main properties of the degree, which are applied in Section 2.2 to obtain the exact number of solutions of (1.1) for some values of β . Section 2.3 includes the gluing information (written by hand), and some applications, and Section 2.4 treats the case without bifurcation. The existence of two different duals will give other solutions. Section 2.5 gives more global bifurcation theorems for positive solutions. Section 3.1 contains a brief survey of the basic properties of the Liénard index, and in Section 3.2 a global maximum theorem for positive solutions (i.e. each, together with a global minimum, in prior existence of positive solutions of a system arising in chemical reactions).

The different methods are discussed in Chapter 4. The method example (1.1) is extended, this time by using a continuation method based on a local Lyapunov theorem for gradient-dynamical systems. A few variants of (1.2) are also considered. Some indications about the application of similar notions to the study of "singular points" in the case without bifurcation are given in Section 4.2. Finally, in Section 4.3 we use variational methods to improve under additional assumptions the results in Section 3.2.

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3. THE METHOD OF THE NON-UNIFORMITY AND APPLICATIONS

This chapter is mainly concerned with the method of sub and super-solutions and some of its applications.

The first paragraph treats the case of a nonlinear elliptic equation. Here it can be seen that the method is very well-known - we find comments in [19] that there will still details are only because they are the best illustration for the general case of systems, see also [10]. As it was mentioned in the Introduction, existence and multiplicity results for some systems can be obtained by reformulating the problem in such a way that they become well-poseable like the minimax eigenvalue problem we consider in Section 4.2. For this problem we study existence and uniqueness of positive solutions together with some interesting additional qualitative properties. Nextly, concerning the uniqueness of the branch of positive solutions we obtain. Finally, the last paragraph considers the extension of the method of sub and super-solutions to systems, and includes a rather general existence theorem. Some applications are also given.

3.1. AN EIGENVALUE PROBLEM FOR NON-LINEAR ELLIPTIC EQUATIONS

Our main interest here is to show how the method of sub and super-solutions work in a concrete but very illustrative example, namely the case of a radial nonlinear elliptic equation. This section, which is in some sense continuation of Section 6, methods for solving differential problems by using sub-super-solutions, not only give an existence proof but also provide an alternative scheme to approximate solutions by sequences converging monotonically to those solutions.

Now we consider the nonlinear elliptic problem

$$\begin{aligned} & -\Delta u = f(x,u) \quad \text{in } \Omega \\ & u = 0 \quad \text{on } \partial\Omega, \end{aligned}$$

where

where $\Sigma \in \mathbb{R}$, as in (1) which for $\Sigma = 0$ is a connected bounded open set or $\Omega \times (0,1)$ with a very smooth boundary $\partial\Omega$; in particular, we assume that Ω is $C^{1,\alpha}$, $0 < \alpha < 1$, but the reader may often work with Ω having C^2 or even C^∞ . We also assume that the function $\varrho : \Sigma \times \mathbb{R} \rightarrow \mathbb{R}$ satisfies

$$(1.2) \quad \varrho \in C^3(\overline{\Omega} \times \mathbb{R}),$$

(1.3) For any $y \in \mathbb{R} \times \mathbb{R}^m$, ϱ is increasing in y .

Remark 1.1: The assumption (1.3) can be weakened. It is sufficient that there exists $R > 0$ such that $|x_0 - x_1| \leq R \Rightarrow \varrho(x_0) \leq \varrho(x_1)$. Indeed, in this problem (1.1), (1.2) are equivalent to the problem

$$\begin{cases} u_t + \nabla u \cdot \nabla \varphi(x,t) = 0 & \text{in } \Sigma \\ u = 0 & \text{on } \partial\Sigma \end{cases}$$

where $\tilde{\varphi}(x,t) = \varphi(x,t) + R$ is monotone (1.2), (1.3). Moreover, we will prove below that (1.3) should only be satisfied on some bounded interval $[0, u_+]$ and then (1.3) will follow from (1.2) (cf. Remark 1.5). On the other hand, (1.3) can also be replaced (cf. [12], [13]) by the more general condition that ϱ is locally C^1 , $0 < \alpha < 1$, and there is a $R > 0$ such that (1.4) is satisfied for $\varrho(x,t) \leq R$ instead of ϱ .

Finally, we give the definition of sub and super solutions for our problem.

Definition: A sufficiently smooth function $u_0 \in L^2(\Omega) \cap C^2(\bar{\Omega}) \cap C^1(\bar{\Omega})$ (resp. φ_0) is called a **subsolution** (resp. a **supersolution**) of (1.1) if (1.5) is true.

$$(1.5) \quad u_0(t) + \varphi_0(x,t) \leq 0 \quad \text{a.s. in } \Sigma.$$

$$(1.6) \quad u_0(t) + \varphi_0(x,t) \geq 0 \quad \text{a.s. in } \Sigma.$$

where inequalities are in the ordered positive.

Our aim is to prove the following existence result.

Theorem 1.1: Let $u_0 \in L^2(\Omega)$, $\varphi_0 \in C^1(\bar{\Omega})$ be a subfunction (1.5), a supersolution of (1.1), (1.6) such that $|u_0| \leq M$ in Σ . Then there exists u and φ solutions of (1.1), (1.2) satisfying $u_0 \leq u \leq \varphi_0$. Moreover, if φ_0 is a maximal (resp. minimal) in the sense that $\varphi_0 \leq \varphi$ (resp. $\varphi \leq \varphi_0$) then $u \leq \varphi$ (resp. $\varphi \leq u$).

A few remarks seem to be in order before giving the proof of theorem 1.1. First, this is not the most general version of this kind of existence results. It is given the constants M to be sharp. On each occasion here it is to demonstrate how the method works, and it is always conceivable from this point of view to get rid of important technical difficulties. In fact, the theory can be extended to much more general class of "obstacle" problems in general Banach spaces (cf. [6], [17]).

Firstly, let us point out that both the initial and the maximal solutions u and φ may oscillate, and this is obviously the case when the solution is unique. (By this we mean solution which is classical in Σ , i.e., a function in $C^1(\bar{\Omega}) \cap C^2(\bar{\Omega})$, but now, necessarily has two different extrema, i.e., p.). Banach's fixed point theorem, Theorem 1.3 has no implications concerning uniqueness, which should be proved. (Note that there is a by a different account, requiring the particular features of the corresponding problem, cf. [1] in the \mathbb{R}^d -case).

On the other hand, this method does not provide all the properties of (1.1), (1.2). One way to make this is to add the an additional weak term or more sublinear, but there are also quite general, and intense, regularity results due to Gutiérrez [1981], [1982] and Ambro [1991] among others. The solution obtained in Theorem 1.3 is universal: there is a weaker, energetic approach, that an unstable motion must be obtained by using our iteration process. Cf. Definition.

Remark 1.2: The fact that u_0 is a subfunction and φ_0 a supersolution does not imply $u_0 \leq \varphi_0$, as it is obvious that any solution is a subfunction (and a supersolution), any problem having more than one solution shows that this is not the case. But it is not difficult

(cf., e.g., below paragraph 1.2), to find examples of strict subordinations where α (only) is strict superordination. Here strict means strict inclusion according to (1.1)–(1.6).

Remark 1.1: The statement is false for a subordination η_0 and a superordination η , such that $\eta = \eta_0$. As noted the counterexample to [26, pp. 852–854]. However, in this case, the problem has no solution at all.

Theorem 1.1 is proved by reducing the problem of finding solutions of (1.1)–(1.2) to the equivalent problem of solving the equations at fixed points of a suitable nonlinear operator T , defined on a function space adequately chosen. What we intend to show is that in this case we should take care not only of the topological (continuity, compactness) properties of T , but also use the fact that T preserves the natural order in the corresponding function space.

Now we define one nonlinear operator T in such a way that its fixed points are (real-valued) solutions of (1.1)–(1.2). The idea is

$$\begin{aligned} T : C^{1,0}(\Omega) &\longrightarrow C^{1,0}(\Omega) \\ z &\longmapsto \eta_0 \end{aligned}$$

where, for $u \in C^{1,0}(\Omega)$, η_0 is the (unique) solution of the linear problem

$$\begin{aligned} T : C^{1,0}(\Omega) &\longrightarrow C^{1,0}(\Omega) \\ z &\longmapsto \eta_0 \end{aligned}$$

(1.7) $\quad \begin{aligned} &+ \Delta u + f(u) = 0 \quad \text{in } \Omega \\ &u = 0 \quad \text{on } \partial\Omega \end{aligned}$

(we apply the usual notations for the spaces of Hölder continuous functions, $C^{k,\alpha}(\Omega)$, $k \in \mathbb{N}$, $0 < \alpha < 1$, and Banach spaces $L^p(\Omega)$, $1 \leq p \leq \infty$, see the corresponding formal). We also use the space $C^1(\Omega)$ and its norm $\| \cdot \|_{C^1(\Omega)}$.

First, we observe T is well-defined. Indeed, if $u \in C^{1,0}(\Omega)$, $\|u\|_{C^{1,0}(\Omega)} < C_0$, then by the well-known C^0 -theory for linear operators

(Browder's theory) (see, e.g., [20, 1, 75]) there is a unique solution to $+ \Delta v + g(v) = 0$ in Ω (cf. (1.7)–(1.8)). Moreover, it is an easy task to prove, by using Schauder's estimates, that T is compact, i.e., that T is continuous and that, due to boundedness, $T(\bar{\Omega})$ is relatively compact.

We shall return later to the links between strict, compactness and the choice of the function spaces for the bounds. Let us say only that, for this problem, it is still possible to work in $C_0(\bar{\Omega})$, $C^1(\bar{\Omega})$, or even $L^2(\bar{\Omega})$.

The following results, which follow easily from the Maximum Principle, show that T between “only” with respect to the natural order in the function space, there is a more “nice” $\eta_0 > 0$ for any $u \in \Omega$.

Lemma 1.1: η_0 is order-preserving. I.e., $u \geq v$ implies $Tu \geq Tv$.

Proof: From (1.7)–(1.8) that u and v follows by (1.8)

$$-\Delta u - g(u) = f(u) \geq f(v) = -\Delta v - g(v) \geq 0 \quad \text{in } \Omega.$$

Since $\eta_0 = 0$ on $\partial\Omega$,

$$\text{and by the Harnack Principle, } Tu - Tv \geq 0.$$

and by the definition of T η_0 is the solution of

$$\begin{aligned} &- \Delta \eta_0 - f(\eta_0) = 0 \quad \text{in } \Omega \\ &\eta_0 = 0 \quad \text{on } \partial\Omega \end{aligned}$$

and thus together with (1.8) (1.8) gives

$$\begin{aligned} &- \Delta \eta_0 - f(\eta_0) = 0 \geq -f(\eta_0) + g(\eta_0) \geq 0 \quad \text{in } \Omega \\ &\eta_0 = 0 \quad \text{on } \partial\Omega \end{aligned}$$

and with by the maximum principle, $\eta_0 = 0$ in Ω .

Lemma 1.2: If u is a superordination, then $\eta_0 = Tu$.

Indeed,

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Using of Theorem 1, we define by \mathcal{I}_n the following measure

It is interesting to note that the time constant of the first-order rate process is $\tau = \ln 2 / k_1 = 1.38 \times 10^{-3} \text{ s}$, which is considerably longer than the time constant of the second-order rate process ($\tau = 1.38 \times 10^{-4} \text{ s}$). This indicates that the second-order rate process is dominant.

(usually they converge to $\langle \psi^A, \psi^B \rangle$) for any $A, B \in \{1, \dots, N\}$. On the other hand, the momenta $p_{AB} = \langle \hat{P}_A, \hat{P}_B \rangle$ implement the CCRs between the two variables. Since \hat{P} is continuous,

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$$= \frac{(\alpha)}{\alpha} \cdot \frac{1}{\alpha} = \frac{(\alpha)}{\alpha^2} \Rightarrow \alpha \neq 0 \quad (\alpha \in \mathbb{R})$$

What is known (or at all) which follows. If not otherwise noted, a

Table 11. Summary of the results of the model calibration.

$$n \geq \left\lceil \ln d^N \over \ln e^{-\lambda} \right\rceil + \left\lceil \ln d^N \over \ln e^{-\lambda} \right\rceil - 1 - O(\sqrt{n}) = \left(\left\lceil \ln d^N \over \ln e^{-\lambda} \right\rceil - 1 \right) \ln d^N.$$

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¹ See also the discussion by Hirschman (1990) on the role of culture in political development.

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are very reasonable for $\alpha \geq 1$. In fact, we can assume $\alpha = 1$ in the following discussion. This clarifies the above notation. Hence (1.4) has no surprises. We can also use the notation in Remark 1.1 by noticing that we do not need to assume E irreducible.

This makes a big difference with respect to the case of systems (6), Section 4.3:

It is clear that the method fails, at an essential way, on the nonlinear problems (the one also linear problems) and on the numerical C^0 (isoparametric theory) cf. [18], [75]) and C^1 (hyper-possibly-bijective) [13] (isoparametric results on the paper). This implies that in the case of elliptic equations, the method is forced to switch under differentiable operators, roughly speaking. Cf. the Remarks for some of the possible extensions of the method.

1.3. PARTITION SUGGESTIONS FOR A NONLINEAR EIGENVALUE PROBLEM

In this section it is shown how the ideas of the section of sub and superimposition may be applied to the study of the existence of positive solutions of the nonlinear eigenvalue problem

$$(1.13) \quad -\Delta u + \varphi(u) = \lambda u \quad \text{in } \Omega$$

$$(1.14) \quad u = 0 \quad \text{on } \partial\Omega.$$

where Ω is a smooth bounded domain in \mathbb{R}^n , λ is a real parameter, and $\varphi : \mathbb{R} \rightarrow \mathbb{R}$ satisfies the assumptions

$$(1.15) \quad \varphi \in C^2, \quad \text{increasing, and, } \varphi(0) = \varphi'(0) = 0.$$

(1.14) is strictly increasing for $u > 0$ and strictly decreasing for $u < 0$:

$$(1.16) \quad \begin{cases} \frac{d\varphi}{du} > 0 & \text{for } u > 0 \\ \frac{d\varphi}{du} < 0 & \text{for } u < 0 \end{cases}$$

problems of this kind arises in the theory of nonlinear resonance (cf. [28], [108]) and have been studied by several authors (cf. [18], [188], [187], [120], [131]).

Firstly, let us introduce a very useful notation. If the function φ is smooth enough (this means C^5 or even C^∞) the eigenvalues λ smooth enough (this means C^5 or even C^∞) the eigenvalues

$\lambda_1 < \lambda_2 < \lambda_3 < \dots$

$$(1.17) \quad \lambda = \lambda(u) \quad \text{on } \Omega.$$

depending on u (a multiplicity, going to a limit $u \rightarrow +\infty$ and depending in a decreasing monotonic way on $u \geq 0$ (cf. [16], [43], [106]).

The classical book by Deimling-Krasnosel'skij [4] is a good reference for all which concerns the properties of operators and eigenfunctions. In particular the maximum and minimum dependence on the coefficient, μ and the domain Ω and the variational characterization of the eigenvalues, etc., which will be very useful in the following.

In the particular case $\varphi \equiv 0$, we will use the notation $\lambda_n = \lambda_n(\Omega)$ for the eigenvalues of

$$(1.18) \quad -\Delta u = \lambda u \quad \text{in } \Omega, \quad u = 0 \quad \text{on } \partial\Omega.$$

It is well known that $\lambda_1(\Omega)$ is simple, i.e., it has no multiple ones, and the corresponding eigenfunction does not change sign in Ω . For $\lambda \geq 0$ we will denote by Ψ_λ the expectation of λ , such that $\Psi_\lambda = 0$ and $|\Psi_\lambda|_* = 1$ (normalization condition). Hence we have

$$(1.19) \quad \begin{cases} \Psi_\lambda > 0 & \text{if } \lambda > \lambda_1 \\ \Psi_\lambda = 0 & \text{if } \lambda = \lambda_1 \\ \Psi_\lambda < 0 & \text{if } \lambda < \lambda_1 \end{cases}$$

$$(1.20) \quad \begin{cases} -\Delta \Psi_\lambda = \lambda \Psi_\lambda & \text{in } \Omega \\ \Psi_\lambda = 0 & \text{on } \partial\Omega \end{cases}$$

A preliminary result concerning the behavior of solutions of (1.14), (1.18), (1.19), (1.20) takes the ground by means a comparison argument.

Lemma 1.4 (problem (1.14)-(1.18) has no nontrivial solutions for $\lambda \in \Lambda$, $\varphi \in \mathcal{C}^5$) $\lambda \in \Lambda$, $\varphi \in \mathcal{C}^5$ problem (1.14)-(1.18) has no nontrivial solutions for $\lambda \in \Lambda$, $\varphi \in \mathcal{C}^5$. Because u is a nontrivial solution, we can define the

- (1.22) $\begin{cases} \frac{\partial u(x)}{\partial x} & \text{for } u(x) \neq 0 \\ 0 & \text{for } u(x) = 0 \end{cases}$
- Notice that this function will be denoted by $\mathbb{I}_{\{u(x) \neq 0\}}$, i.e., in other words, it is equal to one if $u(x) \neq 0$ and zero if $u(x) = 0$.

Finally, we show that h is continuous and (1.42) provides $h_0 > h_1 \geq 0$.

Indeed, this function will be denoted by $\mathbb{I}_{\{u(x) \neq 0\}}$, i.e., in other words, it is equal to one if $u(x) \neq 0$ and zero if $u(x) = 0$.

Finally, we show that h is continuous and (1.42) provides $h_0 > h_1 \geq 0$.

Indeed, this function will be denoted by $\mathbb{I}_{\{u(x) \neq 0\}}$, i.e., in other words, it is equal to one if $u(x) \neq 0$ and zero if $u(x) = 0$.

And by the above and (1.42), we get that $h_0 > h_1 \geq 0$.

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- PROOF.** The approach follows immediately from Theorem 1.1 and the observation that $\nabla u = 0$ if and only if u is constant. \square

THEOREM 1.3. Problem (1.11)-(1.12) has at least a maximal positive solution in Ω . \square

PROOF. The approach follows immediately from Theorem 1.1 and the observation that $\nabla u = 0$ if and only if u is constant. \square

THEOREM 1.4. The solution u of problem (1.11)-(1.12) is unique up to a constant multiple.

PROOF. The proof follows immediately from Theorem 1.3.

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strictly increasing for $u > 0$. Then $\theta = \pi$, but the second part, involving odd k -periodic solutions, does not work if they are not available. We will return to this problem in Section 3.5.

Remark 3.13. The same method can be applied to (3.15), in particular by the assumption that f is asymptotically linear (cf. (2.6) below).

The main difference in this case is that a different trick should be applied to the equations (cf. (63)).

We have obtained some results concerning existence and uniqueness of maximal positive solutions of (3.13)–(3.15) depending on the values of the real parameter λ ; it is possible to obtain, in exactly the same way, the corresponding results for maximal negative solutions.

The next propositions provide some complementary results concerning qualitative properties of the "branch" of positive solutions. These results will be contained again in three notes (cf. in particular

Section 3.4). In fact, the following proposition says that iff (3.13) is the only nontrivial positive solution for the value λ of the parameter, then ψ_{λ} is an increasing continuous "branch".

Proposition 3.14. The mapping $\lambda \mapsto \psi_{\lambda}$ from $(3.7) \cap$ (note (C)) to \mathbb{R}_{+} is continuous. Moreover, if $\lambda_1 > \lambda_2$, then $\psi_{\lambda_1} > \psi_{\lambda_2}$ (order plan).

Proof. It is an easy exercise by using sub and superolutions and the continuity of F , cf., e.g., [61], [20].

3. Note addition.

Proof. If $\lambda = 0$, by assumption (v)-(ii), it follows that the function $u = u(t)$ is unique, and then the uniqueness is an easy consequence of the relation in (17).

Remark 3.14. The first part of those 1 works equally if $\lambda \neq 0$ and φ are two ordered positive solutions (i.e., $\varphi \geq \varphi$ for $t \in [0, 1]$ and φ is a λ -solution, instead of (3.14), which means strict inequality, a strict inequality assumption, namely (2.6) periodic decreasing for $u > 0$ and

and by (3.14), and the continuity of u (in t) $\varphi < u < \varphi$). Suppose now that λ and φ are such that $u = u_0 < \varphi < u$. By applying theorem 1.3 with $\lambda_1 = 0$ and the other assumptions as before, we prove the existence of a nontrivial solution u with $0 < u_0 < u < \varphi$. But note that the fixed point of the map $u \mapsto \varphi$ is

$$(3.28) \quad \varphi(u) = \varphi(\varphi(u)) = \int_0^1 F(u-t)\varphi(u-t)dt = \int_0^1 u\varphi(u-\frac{t}{u})dt = \frac{\varphi(u)}{u},$$

and by (3.14), and the continuity of u (in t), $u < \varphi$.

Remark 3.15. The same method can be applied to (3.13)–(3.15), in particular by the assumption that f is asymptotically linear (cf. (2.6) below).

Remark 3.16. This is by no means the best smoothness result for the branch ψ_{λ} . We will show below (Section 3.5) that ψ_{λ} is actually C^1 as a consequence of (3.13)–(3.15). On the other hand, the monotonicity result can be made more precise by using the maximum principle.

Remark 3.17. This is also possible to prove that ψ_{λ} is the only discontinuous point for positive solutions, cf. (1.13)–(1.16). On the other hand, another useful result is contained in the following proposition.

Proof. Which uses sub and superolutions and comparison arguments.

on we find in (4.1)

EQUATION 4.2 $\frac{d}{dt} u + \lambda u = 0$, $u(0) = 1$, $u(t) = e^{-\lambda t}$.

$$(4.27) \quad y(t) = \begin{cases} 0 & \text{if } t < 0, \\ e^{-(\lambda - \mu)t} & \text{if } t \geq 0. \end{cases}$$

Finally, a more simple comparison argument (cf. [26]) shows that:

in the interval $t_1 < t < t_2$, the three solutions we know, namely the trivial solution, the unique positive solution and the unique negative solution, are the only ones. A more complicated proof of this result was given before in [13] by using topological methods (cf. Section 3.2).

(4) case (γ, ζ) .

EQUATION 4.28 $\frac{d}{dt} u + \lambda u = \gamma u^2$. Thus problem (4.1)-(4.2) has exactly 3 solutions.

PROOF If u is a nontrivial solution, then 0 is either

$$\begin{aligned} &= \lim_{t \rightarrow 0^+} \frac{1}{u(t)} u' = \gamma u_0^2 \quad \text{by (4)} \\ &\text{Hence } \lambda = \frac{1}{2} \left(\frac{1}{u_0^2} - 1 \right) \text{ for some } 2 \leq \lambda < \infty. \text{ But } \frac{du}{dt} = 0, \text{ and this gives} \\ &\frac{1}{2} \gamma = 1 = \frac{1}{2} \left(\frac{1}{u_0^2} - 1 \right) = \frac{1}{2} (0) - 1. \end{aligned}$$

Hence $\lambda = 1$, and $u > 0$ or $u < 0$.

The results in this section can be summarized in the following diagram:

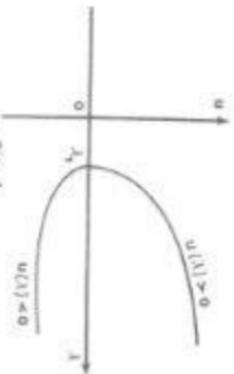


Fig. 1

1.3. THE METHOD OF LINER AND SUPERPOSITION FOR FURTHER APPLICATION

We begin in this section the analysis of solutions for some nonautonomous systems. For the sake of simplicity we only treat the case of systems with the equations $\dot{x}_1 = h_1(t)x_1$, $\dot{x}_2 = h_2(t)x_2$.

We recall in this section the relations of stability for some nonautonomous systems. For the sake of simplicity we only treat the case of systems with the equations $\dot{x}_1 = h_1(t)x_1$, $\dot{x}_2 = h_2(t)x_2$.

EQUATION 4.29 $\dot{x}_1 = g_1(t)x_1$, $\dot{x}_2 = g_2(t)x_2$.

PROOF $x_1 = q_1(t)$, $x_2 = q_2(t)$

$\dot{x}_1 = q_1' = g_1 q_1$, $\dot{x}_2 = q_2' = g_2 q_2$.

where q_1 is a smooth bounded solution in \mathbb{R} and the nonlinear terms g_1 and g_2 satisfy the smoothness assumption

$$(4.31) \quad \dot{g}_1, \dot{g}_2 \in C^1.$$

PROOF $\dot{x}_1 = g_1 q_1$, $\dot{x}_2 = g_2 q_2$.

Now it seems natural when applying as usual the method of sub and superpositions to systems, to add some monotonicity assumptions on the growth of g_1 and g_2 as function of q_1 and q_2 .

The first (natural) reasonable extension is to assume that g_1 and g_2 are both, respectively, increasing in the variable q , and vice versa. In this case, if we adopt an (equivalent) monotone definition of nonautonomous systems for systems, the following lemma would be true (see [1, 29]-[1, 36]). More precisely, if q and 0 satisfy (4.31) and are independent variables in q and q , and if we denote a solution (q_1, q_2) and a superposition (q_1^*, q_2^*) of (4.29) by (q_1^*, q_2^*) we get a new smooth function q^* satisfying

$$(4.32) \quad \begin{aligned} &= \dot{g}_1 q^* = g_1(q_1^*, q_2^*) = 0 = -\dot{g}_2 q^* = g_2(q_1^*, q_2^*) \\ &\text{in } \mathbb{R}. \end{aligned}$$

Since, i.e. $q_1^* \geq 0$, $q_2^* \geq 0$ is the corresponding version of Theorem 1.1

in time and its phase can be varied are in a completely similar way.

Moreover, it is always possible to make F (resp. Φ) increasing.

Let us consider the generalized (cf. remark 1.3) the adjoint problem

$$(3.26) \quad \begin{aligned} & u_{tt} + u_{yy} = f(u,y) + g_0, \quad t \in [0,T], \\ & u|_{t=0} = u_0, \quad u|_{t=T} = u_T, \end{aligned}$$

$$(3.27) \quad \begin{aligned} & u_{yy} + u_{tt} = g(u,y) + h_0, \quad t \in [0,T], \\ & u|_{y=0} = u_0, \quad u|_{y=\bar{y}} = u_{\bar{y}}, \end{aligned}$$

where $R = 8$ is much more than the right-hand side in (3.25) (resp. in (3.26)) is increasing in y (resp. in t).

However, if f is not increasing in y , or if g is not decreasing in y , then we claim the minimum operator T defined in the proof of Theorem 3.3 by means of the system (3.26)-(3.27) is no more symmetric.

Indeed, since in the terms for the closure of T the monotonicity does not work, there is no reason for the closure of T to be non-decreasing. Moreover, if f is linear, it is interesting in y (resp. in t), or, more generally, if each nonlinear term is increasing in the "differentiable variables", the system is called **strict-monotone**. There is a large literature on this subject.

Another possibility is to use, when monotonicity properties of F and Φ allow it, different terms instead of weaker, particular interesting estimates, etc., to obtain monotone sequences corresponding to natural and natural solutions (cf. e.g. [179], [170], [161]). On the other side, if f and g do not satisfy natural monotonicity assumptions with the above definition of u and u_{yy} in (3.26)-(3.27) the terms entering the system (3.1) in Euler form, as above the Ellington numbers, the system we consider is

$$\begin{aligned} & u_{yy} - u_{tt} - u_{yy} + u_{tt} = f(u,y) + g_0, \quad t \in [0,T], \\ & u|_{t=0} = u_0, \quad u|_{t=T} = u_T, \\ & u|_{y=0} = u_0, \quad u|_{y=\bar{y}} = u_{\bar{y}}, \end{aligned}$$

It is easy to see that, for this system (3.26)-(3.27) (E), it is a monotone-

like and $(u_0, u_T) \in (\mathcal{O}_0, \mathcal{O}_T)$ is a representation. Hence, if the theorem is well-known, it is not true.

In this well-known case, there is a solution u_0^{ref} with $u_0^{\text{ref}} \geq 0$ and $u_T^{\text{ref}} \geq 0$. But if $u_0^{\text{ref}} \leq 0$, then $\frac{d}{dt} u_0^{\text{ref}} \geq 0$ on the time interval $[0, T]$.

This counterexample seems to indicate that the above definition of sub and superlution (3.26)-(3.27) is not valid, at least without additional nondecreasing assumption. As this about how to modify the definition of sub and superlution can be observed by looking, from a simple different point of view, at the proof of Theorem 3.3.

Indeed, by the other-growing properties of the operator T (Lemma 3.4.3) it follows immediately that the interval $I = [0, T]$ is followed (immediately) with the interval $I = [0, T]$ by $I = [0, T]$ and $I = [0, T]$ by $I = [0, T]$ (cf. (3.26)).

$$B = \lim_{n \rightarrow \infty} \sigma_n^D = (\mathbf{I} - \frac{1}{n} \partial_y) \times \mathbf{a}(x) \in \mathcal{C}^0([0, T]; \mathbb{R})$$

is invariant by \mathbf{T}_t (resp. \mathbf{T}_y) ($\mathbf{T}_y = \mathbf{I}$), and this requires to use Φ (3.5) + Definition 3.3. Then positive elements of the nonincreasing assumption on g and h are satisfied, of course, this result would be weaker than Theorem 3.3. Here the application to systems could be interesting.

We remark that, even if the result is weaker, it requires some additional assumptions on the function space we employ. The proof of Theorem 3.3 was carried out in $C^1([0, T])$, but many other function spaces, as we have pointed out, were usually well-adapted, e.g. $C^1([0, T] \times \mathbb{R}^2)$ or $L^2(\Omega)$. To apply Debruyne's fixed point theorem we need a compact operator \mathbf{P} and a closed, bounded, convex subset \mathcal{K} . There is a well-known problem with \mathbf{P} , but the function Φ is not bounded in $C^1([0, T] \times \mathbb{R}^2)$ (and even in $C^2([0, T] \times \mathbb{R}^2)$, cf. e.g. [170, p. 427]), while it is in $L^2(\Omega)$ or $L^2(\Omega) \times L^2(\Omega)$. The use of these spaces would imply, as usual, some auxiliary arguments to prove that the weak solution we obtained are actually closed (nonempty). The "cure" for these difficulties is that the norm involves two derivatives in most cases, while the other involves only the function. But, with a suitable choice of the function space and the definition of sub and superlutions, it is possible to prove an estimate allowing for Φ to be noncontinuous (Theorem 3.3 without nondecreasing assumption on \mathbf{P} and Φ).

DEFINITION: The pair $(u_0^{(0)}, v_0^{(0)}) = (u_0^{(0)}, v_0^{(0)})$ is called a **sub-superposition** for a simple sub-superovolution of system (1.39)-(1.40) if

$$\begin{aligned} u_0^{(0)} - u_0 &= \psi^2(\alpha) + k^{(0)} = k^{(0)} + u_0 + \alpha, \quad \psi = \varphi^0 \quad \text{and} \\ (1.46) \quad \psi &= \varphi^{(0)}(\alpha), \quad \psi = \varphi^{(0)} + k^{(0)}, \quad \psi = \varphi^{(0)} + k^{(0)}. \end{aligned}$$

$$(1.39) \quad u_0^{(0)} = g(u_0^{(0)}) + \psi = x_0^{(0)} + g(u_0^{(0)}), \quad \psi = \varphi^{(0)} + k^{(0)}$$

$$(1.40) \quad v_0^{(0)} = g(v_0^{(0)}) + \psi = y_0^{(0)} + g(v_0^{(0)}), \quad \psi = \varphi^{(0)} + k^{(0)}$$

where we used the notation

$$(\varphi, \psi) = (\varphi + \varphi^0)(\alpha) \cdot (\varphi(\alpha) + \varphi^{(0)}), \quad \alpha = 0, \quad \alpha \in \mathbb{R}.$$

Remark 1.18: It is clear that there is a coupling between the components of a sub-superovolution, contrary to the case of deflation (1.32)-(1.33). However, with certainty when φ and φ^0 are separable, interesting is, is and ψ ? If not, the last is much more difficult than the preceding case (cf. remark 1.17) for the underlying potential interpretation).

Remark 1.19: The idea of this definition comes from some classical problems of physics for systems of arbitrary differential equations. Some different versions of it have appeared in works concerning both geodesic and celestial mechanics-dynamical systems, often in a simplified form due to the nondegeneracy properties of the function $\alpha(\cdot)$, cf. (1.76)-(1.82).

Remark 1.20: If every Lagrangian potential instead of sub-superovolution in the case of anisotropic dynamics is a constant function, this is equivalent to say that the sub-superovolution becomes an α -isotopic isotopy, where the vector field $\dot{\alpha}(t)$ is always pointed to the interior of the rectangle along its boundary. The same quantitative interpretation as the case of definition (1.32)-(1.33) says that while the various subpotentials in the four variants are passed to the right, this notion of isotropic isotopy is very important for the study

of motion-difference systems, cf. (1.55)-(1.61).

Theorem 1.26(1): Suppose that $(u_0^{(0)}, v_0^{(0)})$ is a sub-superovolution of the system (1.39)-(1.40) with $\psi = \varphi^{(0)} + k^{(0)}$ satisfying (1.46), then there exists at least a (classical) solution $(u(t), v(t))$ of the system such that $u_0^{(0)} + \psi = u(t) + \alpha$, $v_0^{(0)} + \psi = v(t) + \alpha$.

PROOF: Let $\mathcal{L} = L_0(\dot{u}, \dot{v})$ and $\mathcal{R} = (u_0^{(0)}, v_0^{(0)}) + (\varphi^{(0)}, \psi)$. Then \mathcal{L} is closed, bounded, convex subset of \mathbb{R}^2 . We define a nonlinear operator

$$\begin{aligned} \tilde{\mathcal{L}}(u, v) &:= \mathcal{L}(u, v) - \mathcal{R} = \\ (1.47) \quad \tilde{\mathcal{L}}(u, v) &:= (u - u_0^{(0)} - \varphi^{(0)}(\alpha), v - v_0^{(0)} - \varphi^{(0)}(\alpha)). \end{aligned}$$

where (u, v) is the unique solution of the linear system

$$(1.48) \quad u - u_0^{(0)} - \varphi^{(0)}(\alpha) = \dot{u}_0^{(0)} + \dot{\varphi}^{(0)}(\alpha) = \dot{u}, \quad v - v_0^{(0)} - \varphi^{(0)}(\alpha) = \dot{v}_0^{(0)} + \dot{\varphi}^{(0)}(\alpha) = \dot{v}.$$

$$(1.49) \quad u = \dot{u} = 0 \quad \text{or} \quad \dot{u} = 0, \quad v = \dot{v} = 0 \quad \text{or} \quad \dot{v} = 0.$$

Let $\alpha = 0$ with this the right-hand side in (1.48) (1000, (1.42)). It follows that $\dot{u} = 0$ (1000, (1.42)).

First, \mathcal{L} is well-defined. Indeed, by the classical linear theory, as $\mathcal{L}(0, 0) = \mathcal{L}(0, 0)$, there is a unique $\psi = \varphi^{(0)}(\alpha)$, the zero $\alpha = 0$ solution of (1.45)-(1.46). The same set $\mathcal{L} = \mathcal{L}(0, 0)$ by the

(functional) regularity theory, it follows in a quite straightforward way that \mathcal{L} is compact.

It remains to prove that $\mathcal{L}(0, 0)$ is \mathcal{R} -stable, that $\mathcal{L}(0, 0) = \mathcal{L}(u_0^{(0)}, v_0^{(0)})$ and $\mathcal{L}(0, 0) = \mathcal{L}(u_0^{(0)}, v_0^{(0)})$. We only show that $\mathcal{L}(0, 0) = \mathcal{L}(u_0^{(0)}, v_0^{(0)})$, the other cases being analogous. We prove in a completely analogous way. By (1.39) with $\psi = \varphi^{(0)}(\alpha)$ and (1.41)

$$\begin{aligned} 0 &= \mathcal{L}(0, 0) = \mathcal{L}(u_0^{(0)}, v_0^{(0)}) + \mathcal{L}(u_0^{(0)}, v_0^{(0)}) + \mathcal{L}(u_0^{(0)}, v_0^{(0)}) \\ &= \mathcal{L}(u_0^{(0)}, v_0^{(0)}) + \mathcal{L}(u_0^{(0)}, v_0^{(0)}) + \mathcal{L}(u_0^{(0)}, v_0^{(0)}) + \mathcal{L}(u_0^{(0)}, v_0^{(0)}). \end{aligned}$$

In particular, if the inequality $\|u\|_{L^{\infty}(\Omega)}^{q_0}$ is small, the same conclusion holds, provided $\alpha \in (1, \frac{1}{2}) \cup (3, 6)$.

In what follows we give an application of Theorem 4.4, contained in [3].

$$\begin{aligned} T &= \int_0^T \left[-\Delta (u(t)) u(t) + \left(1 + \tilde{u}(t)^2 - 4 (u_t(t))^2 + W(u(t), t) \right) u(t) \right] dt \\ &\quad + \int_0^T \left[\frac{\partial}{\partial t} \left(\frac{1}{2} \|u(t)\|^2 \right) + \left(1 + \tilde{u}(t)^2 \right)^2 - \left(\frac{\partial}{\partial t} u(t) \right)^2 \right] dt \\ &= \int_0^T \left[\frac{1}{2} \|u(t)\|^2 + \left(1 + \tilde{u}(t)^2 \right)^2 - \left(\frac{\partial}{\partial t} u(t) \right)^2 \right] dt, \end{aligned}$$

$$u = u(t) = u_0 + \int_0^t u(s) ds.$$

where $\tilde{u} \in \mathbb{R}$, $u_0 \in \text{dom } C^1$. Assume that they satisfy

$$(4.40) \quad \begin{cases} \tilde{u}(0) = 0, & \tilde{u}'(0) = 0, \\ \tilde{u}(T) = 0, & \tilde{u}'(T) = 0, \\ u(0) = u_0, & u'(0) = 0, \end{cases}$$

$$(4.41) \quad \begin{cases} \tilde{u}(0) = 0, & \tilde{u}'(0) = 0, \\ \tilde{u}(T) = 0, & \tilde{u}'(T) = 0, \\ u(0) = u_0, & u'(0) = 0, \end{cases}$$

Making into account that the second integral in positive in monotonicity and the boundary integral is negative, one has used a classical convexity argument, namely that if $x \in \mathbb{R} \backslash \{0\}$, then $x^2 = x \cdot x^2 \geq 0$ and $x^2 = 0 \iff x = 0$. By summing fixed points theorem, there is at least a fixed point of T , which is a weak solution of the system.

Finally, we show that (4.41) is a classical solution. If and only if $\tilde{u} \in \mathbb{R}$ is a fixed point of T , obviously, if $(\tilde{u}, 0)$ is a fixed point, then $\tilde{u} = u^2(0)$, for any $1 \leq p \leq \infty$, and, by Hörmander's lemma, $u = u^2(0)$, for any $1 \leq p \leq \infty$. Hence $F(u) = \theta(u, \tilde{u}) \in C^1(\mathbb{R})$ and by Schauder's theorem, $u = u^2(0)$, for any $0 \leq \tilde{u} \leq 0$

$$(4.42) \quad \text{there is } 0 > \theta > 0 \text{ such that } 0 = F(u) = K \text{ for all } u \in \mathbb{R}.$$

This is exactly mean that $(u^2(0), 0) = (\tilde{u}, 0)$, where $u = u^2(0)$.

$$\begin{aligned} &\rightarrow u = 0 \text{ from both side of } 0 \\ &\Rightarrow u = 0 \end{aligned}$$

and we get a classical solution.

Remark 4.14 The estimation (4.10) in the previous section is not true that case where \tilde{u} and u are not initially Lipschitz and non maximal monotone graphs, but their generalizations, which are induced by two boundary problems resulting $u \in L^2$ in maximal monotone generalized porosity (see Cannarsa et al., [22], [45]).

$$\tilde{u}(t) = u_0 + \int_0^t u(s) ds,$$

by a direct argument, last LT 2 and 4 are giving Lipschitz and the corresponding Lipschitz constants are "small", then the solution is also possible, by using nonlinear analysis of G. Z. Fix and supercavitation [42], to prove immediately that the inclusion and (4.42).

F. TOPOLOGICAL METHODS: THE LERAY-Schauder INDEX,
GLOBAL INFORMATION MEASURING AND APPLICATIONS.

This chapter is devoted to topological methods, in particular the Leray-Schauder degree, and some of their applications to the nonlinear elliptic problems we are studying. The topological degree is, in finite-dimensional spaces we introduced in a very concise paper by Leray-Schauder [77] in 1934, generalizing the homological degree in finite-dimensional sets to Banach:

The Leray-Schauder degree is one of the more general and powerful methods to prove existence theorems for nonlinear equations. Generally, to the method of sub and super-solutions to the linear conjugate, compact operators do not play an important role here; in particular this implies that applications to nonlinear partial differential equations are not limited to second order equations where the Maximum Principle holds. In Laplace's example of this situation see (fourth section) via Karaman's remarks on nonlinear elasticity, where conclusions remain true and available.

The topological degree is a very useful tool to prove existence results for nonlinear problems from this point of view the main difficulty is usually to obtain a better estimate on the solutions in order to be able to apply the theory (in this chapter we also want to "justify" the statement of subsections, obtaining lower bounds, or worse giving the exact number of solutions (cf., Theorem 2.2 below). Moreover, the degree is the main tool to prove the so-called global bifurcation theorem, which are very useful, coupled with a degree argument and simple some additional information, to prove existence theorems. Section 2.3 gives an indirect account of the definition of topological degrees, followed by a list of its main properties. These properties are used in Section 2.4 to compute the exact number of solutions of the nonlinear eigenvalue problem. In Section 2.4 - Section 2.6 besides (without proofs) a global bifurcation theorem by Rabinowitz and Rabinowitz-Lazer [11] is given.

Some applications, in particular to the same problem in Section 4.2, the local minimization, Section 4.5 is devoted to global bifurcation theories for problems involving. Section 5.3 treats the same point, fixed and finally Section 5.7 contains its application to transversalization systems.

2.1. THE TOPOLOGICAL INDEX: MAIN PROPERTIES.

In this section we intend to give only a very general and abstract idea of the definition of the topological degree. Let us start from dimension, and then in the infinite-dimensional case, for a systematic exposition, introducing full proofs, cf. [197], [185], [198], [122].

Let Ω be a bounded open subset of \mathbb{R}^n ($n > 1$), and let $\theta = \frac{\pi}{100} \cdots \frac{\pi}{20}$, $\theta = \mathcal{C}^{-1}(1) \subset (\frac{\pi}{10}, \frac{\pi}{2})$. We define $\alpha = (\theta + 0.1, \theta + 0.1)$, where $\mathcal{C}(\theta)$ denotes the domain at θ at the point $\theta + 0$ (which is called the set of boundary points: $\theta = \frac{\pi}{20}$ if $\theta = 0.05$) is easy to see that $\#^{-1}(\alpha)$ is a discrete integer number, and hence $\#$. Consequently the number

\#(q, 0, 0) = \sum_{\theta \in \alpha} \#(q, \theta, 0)

where $\#$ denotes the right-hand side Lebesgue measure, is well-defined and called the degree of q relative to the point 0 with respect to Ω . In this way the degree can be considered as a kind of "numerical function numbers", followed by a list of its main properties. These properties are used in Section 2.4 to compute the exact number of solutions of the nonlinear eigenvalue problem. In Section 2.4 - Section 2.6 besides (without proofs) a global bifurcation theorem by Rabinowitz and Rabinowitz-Lazer [11] is given.

The first one, if $q(0) = 0$ it is easy to prove by using Sard's theorem that $\#(q, 0, 0) = 1$. The second is more important: we want to define the degree for every continuous map q , then $\#$ is approximately by C^1 functions, and it is necessary to check that

The degree is independent of the approximation.

It is convenient to assume that it is impossible to define there a topology and derive the most obvious map with the same properties.

In finite dimension, indeed, if not, Brouwer's fixed point theorem (which can be proved by using degree theory, cf. e.g., [6.1]) would be valid, which is not the case. Introducing time point topology (or otherwise), we see, in some sense, that the finite simple some computations assumption is violated.

If $\text{dom} \subset E$ and \mathbb{T} is a closed, bounded, convex subset, E should be compact at \mathbb{T} is very unnatural; then it should no longer (and con-

sequently) be necessary, in order to have an extension to infinite dimensions of the topological degree, to work in more abstract sets of the class of all continuous maps involving some compactness conditions: the set of all mappings of the form $\mathbb{T} \times \mathbb{T}$, with \mathbb{T} compact, usually called compact perturbations of the identity or "compact version" (in the terminology of [14]).

Let E be a real Banach space, and let $\mathbb{T} : E \rightarrow E$ where \mathbb{T} is a bounded open subset of E , with \mathbb{T} compact. Then it is possible to define the degree of the map $\mathbb{T} - \mathbb{T}^*$ by using, roughly speaking, the fact that compact operators can be approximated by operators with finite dimensional ranges. Thus, if $b \in \mathbb{T}(E)$, it is possible to define an integer $d(\mathbb{T}, b)$, the Leray-Schauder degree of $\mathbb{T} - \mathbb{T}^*$ relative to the point b with respect to E , with the following properties:

1. Continuity with respect to \mathbb{T} . There exists a neighborhood \mathbb{U} of \mathbb{T} in $L(E, E)$ (space of compact mappings from E into E) with the norm $\|y\| = \deg(\mathbb{T}, E)$ such that for any $\mathbb{T}' \in \mathbb{U}$, $b \notin \text{dom}(\mathbb{T}')$,

$$\deg(\mathbb{T}, b) = \deg(\mathbb{T}', b). \quad (1)$$

(1)

2. If $\mathbb{T}(x) = x$ for all $x \in E$, then $\deg(\mathbb{T}, b) = 1$.
3. The degree is constant on connected components of $E - \text{dom}(\mathbb{T})$.

Then for any $t \in (0, 1)$

$$\deg(\mathbb{T}, t b) = \deg(\mathbb{T}, b).$$

4. Additivity. Let $\mathbb{T} = \mathbb{T}_1 + \mathbb{T}_2$, with \mathbb{T}_1 and \mathbb{T}_2 disjoint bounded open subsets of E . If $b \notin \mathbb{T}_1 \cup \mathbb{T}_2$, then

$$\deg(\mathbb{T}, b) = \deg(\mathbb{T}_1, b) + \deg(\mathbb{T}_2, b).$$

5. Equality. If $d(\mathbb{T}, b) \neq 0$, then there exists a $x_0 \in E$ such that $\mathbb{T}(x_0) = b$.
6. Equality. If $b = 0$ is a fixed and $\mathbb{T} \notin \{0\} \times \text{sign}$:

$$d(\mathbb{T}, 0, b) = d(b, 0 - \mathbb{T}, b).$$

We are now in the position to introduce another topological tool, easily derived from the degree, which will also be useful in the following, the index of local degrees.

Let $\mathbb{T} = \mathbb{T} - \mathbb{T}^*$ as above and let u_0 be an isolated solution of the equation $\mathbb{T}(u) = b$, i.e., a solution such that u is the unique solution in some neighborhood of u_0 . By the successive property

$$d(\mathbb{T}, \mathbb{T}_{u_0}(b), b) = d(\mathbb{T}, \mathbb{T}_{u_0}^{-1}(b), b)$$

for every $0 < r < \mathbb{T}_{u_0}^{-1}(b)$ sufficiently small, and the index of the isolated solution u_0 is defined by

$$i(\mathbb{T}, u_0, b) = d(\mathbb{T}, \mathbb{T}_{u_0}^{-1}(b), b).$$

The following results are very useful, since they allow to calculate the index and then the degree in some situations, not via the

For $\varphi \in V_0$, we reformulate our problem as a fixed point equation

$$(2.8) \quad u = (\varphi + \lambda\varphi + \psi)(u) - \lambda\varphi - \psi$$

We are led to derive the monotonic operator associated with the $\mathcal{L}^{\text{gen}}\text{-fixed point in (2.6)}, \mathcal{L} := \mathcal{L}(\tilde{\Omega}) \subseteq \mathcal{C}(\tilde{\Omega})$, where the $\mathcal{C}(\tilde{\Omega})$ is defined by $\mathcal{P}(u) = \mathcal{L}(u) + \psi(u)$. It is well-known [17]: if \mathcal{L} is continuous and bounded (i.e., \mathcal{L} is bounded and $\mathcal{L}(B)$ bounded), \mathcal{L} is surjective if $\mathcal{L}^{-1}(0) \neq \emptyset$. Then (2.6) is equivalent to

$$(2.9) \quad u = \mathcal{L}(u) + \psi(u)$$

where $\psi = \mathcal{L}^{\text{gen}}: \mathcal{C}(\tilde{\Omega}) \rightarrow \mathcal{C}(\tilde{\Omega})$ is compact, since \mathcal{L} is compact.

The following series of lemmas provide the details (1)-(3) and thus allow the proof of Theorem 2.2.

Lemma 2.3. *There exists $R > 0$ such that if $u \in \mathcal{L}^{\text{gen}}(B_R)$ and $\|u\|_{\mathcal{L}^{\text{gen}}} < R$,*

the preimage $\psi(u)$ is not more difficult than can be found in [14, Lemma 2.1] with $\lambda = 1 - \lambda_2$.

The proof, which is not more difficult, can be found in [14, Lemma 2.1].

(1). By the zero of positive solutions of Section 3.4 above,

$$\begin{aligned} 0 &\leq \int_{\tilde{\Omega}} (|\psi(u)|^2 + (\varphi + \psi(u))\psi(u)^2) \, dx \\ &\leq \int_{\tilde{\Omega}} (|\psi(u)|^2 + |\varphi|^2 + \mathcal{L}^{\text{gen}}(u)\psi(u)^2) \, dx \end{aligned}$$

Integrating by parts and integrating over $\tilde{\Omega}$, we obtain (1)-(2) by Lemma 2.3.

Remark 2.4. Lemma 2.3 proves that, for $\lambda_2 > 0 > \lambda_1^2$, the linear \mathcal{L}^{gen} operator along a nontrivial solution is invertible or, in other words, that these solutions are not deformed.

Lemma 2.5. *There exists $R > 0$ such that $u = \psi(u) \in \mathcal{L}^{\text{gen}}(B_R)$ if and only if $\mathcal{L}^{\text{gen}}(u) = \psi(u)$. Moreover, $\mathcal{L}^{\text{gen}}(B_R) \cap \{u \in \mathcal{L}^{\text{gen}}(B_R) : u \neq 0\} = \emptyset$.*

Proof. The first part is simple (Lemma 2.3). For the second, let $u \in \mathcal{L}^{\text{gen}}(B_R)$ and let φ satisfying $\varphi \in \mathcal{L}^{\text{gen}}(B_R)$. Correspondingly

$$\begin{aligned} 0 &= \mathcal{L}^{\text{gen}}(u) + \psi(u) - u = \mathcal{L}^{\text{gen}}(u) - u \\ &= \mathcal{L}(u) + \psi(u) - u = \mathcal{L}(u) - u \end{aligned}$$

and since it is clear that the estimate for $\varphi = 0$ (Lemma 2.2), implies

and since for $\lambda \in \mathbb{R} \setminus \{0\}$, by the homotopy invariance for the degree

$$\deg(\mathcal{L} + \lambda\varphi, B_R(0), 0) = \deg(\mathcal{L} + \varphi, B_R(0), 0) + \deg(\mathcal{L}, B_R(0), 0) = 1$$

for $\varphi \in \mathcal{L}^{\text{gen}}(B_R)$, this means that \mathcal{L} is not a characteristic value, and for $\varphi = 0$ that no characteristic value is in $\langle 0, 1 \rangle$. Hence $\mathcal{L}(u) - u = 0$.

Let us check (2.6). The estimate $\|\psi(u)\|_{\mathcal{L}^{\text{gen}}} = \|\mathcal{L}^{\text{gen}}(u) - \psi(u)\|_{\mathcal{L}^{\text{gen}}}$ yields the

$$\begin{aligned} &= \|\mathcal{L}(u) - u + \mathcal{L}^{\text{gen}}(u) - \psi(u)\|_{\mathcal{L}^{\text{gen}}} \\ &\leq \|u\|_{\mathcal{L}^{\text{gen}}} + \|\mathcal{L}(u) - u\|_{\mathcal{L}^{\text{gen}}} + \|\mathcal{L}^{\text{gen}}(u) - \psi(u)\|_{\mathcal{L}^{\text{gen}}} \\ &\leq \|u\|_{\mathcal{L}^{\text{gen}}} + \|\mathcal{L}(u) - u\|_{\mathcal{L}^{\text{gen}}} + \|\psi(u)\|_{\mathcal{L}^{\text{gen}}} \end{aligned}$$

Integrating by parts and integrating over $\tilde{\Omega}$, we obtain (2) by Lemma 2.3.

Lemma 2.6. *The solution $u \neq 0$ is isolated and $\deg(\mathcal{L} - u, \mathcal{L}^{\text{gen}}(B_R)) = -1$.*

Proof. As similar way, given $\varphi \in \mathcal{L}^{\text{gen}}(B_R)$, the equation $u = \mathcal{L}^{\text{gen}}(u)$ is equivalent to $\mathcal{L}(u) = \varphi$ in $\tilde{\Omega}$, $u \neq 0$ on $\partial\tilde{\Omega}$. Since $\lambda_1 < 1 < \lambda_2^2$ by Corollary 2.3, $\mathcal{L} - \varphi$ is isolated and $\deg(\mathcal{L} - \varphi, \mathcal{L}^{\text{gen}}(B_R)) = -1$, where φ is the sum of the contributions of the characteristic values in $\mathcal{L}^{\text{gen}}(B_R)$. Then $\mathcal{L} - \varphi$ is the preimage to φ .

and $\lambda = \frac{1}{2} + 1$. Hence $\beta > 1$, $\lambda < 0$ and finally $\lim_{t \rightarrow -\infty} \psi(t, 0, 0) = \lambda$.

2.1. Global bifurcation codimension.

Let θ be a real branch point and let $\theta < 0 < \theta' = \theta + \eta$ (η small). Assume that $\alpha(\lambda) = \lambda - \text{fix}(w)$ is θ -smooth, i.e. all values of the real parameter λ the solution $\psi(\lambda, 0) = \lambda$ which is usually called the global solution; we only consider solutions at the form

$$(2.8) \quad \psi(\lambda, 0) = \lambda + \Omega(\lambda) + o(\lambda)$$

where

$$(2.9) \quad \Omega(\lambda, 0) = (\lambda + \Omega(\lambda))^\alpha$$

with λ and Ω satisfying

(2.10) λ is a compact closure operator.

$$(2.11) \quad 0 < \lambda \text{ compact and } \Omega(0, 0) = 0 \text{ if and only if}$$

λ is bounded, i.e., $\lim_{|\lambda| \rightarrow \infty} \frac{\Omega(\lambda, 0)}{\lambda} = 0$, uniformly for λ on bounded intervals.

In this situation it is well-known that to ensure the point λ_0 for $(\lambda, 0)$ is a bifurcation point of (2.9), a necessary condition is that λ_0 is a characteristic value of λ (cf., e.g., [73]). Very simple counter-examples show that this condition is not sufficient. There are some rather general sufficient conditions: one of them is that λ_0 is a characteristic value of λ (or its integral) multiple of $\text{fix}(w)$ (cf. [33]), but this result has a purely linear character: all that is needed is that there is a unique "small" neighborhood of $(\lambda_0, 0)$ such that every "sub-bifurcated" solution is stable. In fact (3.9) with $\theta \neq 0$, the following theorem, due to Bautin-Witt, shows that the fact that λ_0 has multiplicity has much more

implications concerning in particular the global structure of the solution set of (2.8).

Theorem 2.2 ([63], [72]). Assume (2.8), (2.10), (2.11). Then there exists a compact component C of ψ in the closure $\overline{\Omega} = \overline{\Omega} + \mathbb{R}$ of the set of numerical solutions of (2.8) with that C contains $(\lambda_0, 0)$ and either C is unbounded (then contains $(\lambda_0, 0)$), where λ_0 is a characteristic value of λ , $\lambda_0 \neq 0$.

More precisely, this is the case if the generic, i.e. the standard 1-parameter family of solutions with λ_0 on the one side may have the same degree by homotopy invariance and, on the other, satisfy the assumptions of Corollary 2.2. This means that the index would change from 1 to -1 for some λ to 0 when crossing an eigenvalue of $\text{fix}(w)$, i.e. a coincidence.

It is very easy to see that such cases as Theorem 2.2 actually occur. If $\theta = 0$, the problem is reduced to $\lambda = \lambda_0$, and the non-spectral components are simply the corresponding eigenspaces, all an example of the second (1977) in our book $\theta = \theta^2 = 0$.

which

$$\lambda = \lambda_0 + \lambda \omega \quad \Rightarrow \quad u = \begin{bmatrix} \lambda \\ u_2 \end{bmatrix}$$

$$\lambda = \begin{bmatrix} 1 & 0 \\ 0 & 1/2 \end{bmatrix}, \quad \lambda_0 = \begin{bmatrix} -1/2 \\ 0 \end{bmatrix}$$

is the characteristic value λ_0 of λ , both simple, which are fixed points. A very simple calculation shows that the set of non-spectral solutions is a closed loop in the interval $\lambda \in \mathbb{R}$.

Remark 2.3. Similar results can be proved for bifurcation at infinity, cf. [55].

The main difficulty in order to apply Theorem 2.2 is to know which of the two above possibilities actually occurs. In this very interesting case (a class of nonlinear eigenvalue problems and a class

of quadrilateral element problems). It is *impossible* to show (cf. (3.1), (4.2), [17]) that the continua of solutions bifurcating from simple eigenvalues (as is the case for all the bifurcations for hyperbolicity problems and the first, for some elliptic problems) are bounded. One proof applies largely some qualitative properties of the material. More generally, in the case of transversal problems, the global properties of the eigenfunctions if \mathbf{q}_0 is an eigenfunction for the eigenvalue λ_0 , then \mathbf{q}_0 has precisely (6.1) simple nodes in the corresponding interval. The same properties imply that continua arising from different eigenvalues cannot intersect. In the case of quasi-linear elliptic problems having a "bifurcating" eigenvalue problem at $\mathbf{q} = \mathbf{q}_0$, (4.6), (4.7), both arguments may apply to the first approximation, which is simple and has a position-dependent.

Global bifurcation theories can be used to obtain existence results. For (2.1) — however, no such set exists, even if it is possible to prove that the dimension of solutions grows by Rabinowitz's theorem as "infinity". The bifurcation in the solutions are in fact very sparse. Indeed, all we know in principle about \mathcal{C} is that \mathcal{C} is a closed curve in \mathbb{R}^n (more precisely, there is a very convenient way to get an more additional information as possible in subsection 4).

A derivative step in this direction is to have a global estimation for the solutions at (2.8), if there is a continuous function $\psi: \mathbb{R} \rightarrow \mathbb{R}^n$ such that $\psi(\lambda_0) = \mathbf{q}_0$ (see [4], [11], [13]), and ψ is unbounded, then \mathcal{C} can "go to infinity" "to the right" and/or "to the left", but not "vertically" (cf. Fig. 2).

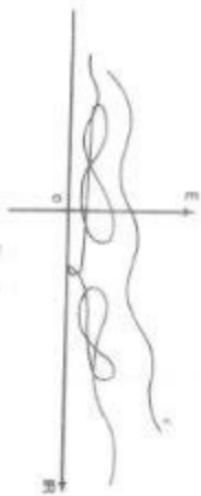


FIG. 2

However, if some more information on solutions is available, then the structure of the solution set can be made more precise. Cf. Sections 4.2 and 5.1, let us apply this result to our problem (2.1), (2.2) with assumptions (2.3)-(2.5). The problem can be written as

$$\mathbf{u} = (\mathbf{u}_\alpha - \mathbf{L}\mathbf{f}(\mathbf{u}))$$

where $\mathbf{u} = \mathbf{L}^{-1}\mathbf{f}^\ast(\mathbf{u})$, since $\mathbf{f}(0) = \mathbf{f}^\ast(0) = 0$ is the origin. We check that for any "unstable" (say, $\mathbf{Q}(0) = \mathcal{C}(0)$, etc.) $\|\mathbf{L}f(\mathbf{u})\| \rightarrow 0$ as $\mathbf{u} \rightarrow \mathbf{0}$ and the linearized problem $\mathbf{u} = \mathbf{L}\mathbf{u}$ has the single eigenvalue λ_0 . Now it is possible to apply Theorem 3.3 (10) and to prove that the component \mathcal{C} is unbounded. On the other hand, even if we have only some more information on solutions of (2.1), (2.2), at our disposal, for $\mathbf{u} \in \mathcal{C}$, where there is only one trivial solution (Lemma 4.4) and there are $\mathcal{S}(\mathbf{u})$ -elliptic extensions (Lemma 3.2), there may still be some "go to infinity on the right", but anyway, even using these supplementary results, the method of sub and superfunctions in Section 4.2 and the minimization method of Section 2.9 give much more information.

These considerations concerning our basic problem are still valid in a rather general setting. When applicable, this method of sub and superfunctions seems to be preferable to global bifurcation arguments in order to prove existence results for nonlinear problems of the form (2.8), because to prove a nontrivial global branching somewhere it seems they allow to get both generic information on the structure of the solution sets. But if sub and superfunctions are not applicable, then topological methods and in particular global bifurcation theorems can be used to prove existence results.

2.4. BIFURCATION OF CONTINUUM IN THE CASE WITHOUT BIFURCATION

Let \mathbf{E} be a real Banach space and let $\mathbf{q}: \mathbf{E} \times \mathbb{R} \rightarrow \mathbb{R}^n$ compact such that $\mathbf{q}(0,0) = 0$ for every $\mathbf{q} \in \mathbf{E}$. Hence $(0,0)$ is the only

(2.12) $\|u - \tilde{u}\|_{L^2(\Omega)} = 0$

meaning in $\mathbb{R}^n \times \mathbb{R}$. The following theorem is an example of the kind of results which can be proven in the semi-bifurcation case.

Theorem 2.8 ([37]). With the above assumptions, if C is the semistable component of the solution set of (2.12) containing $(0,0)$, then $0 = C^0 \cup C^+$, where $C^0 \subset \mathbb{R} \times \mathbb{R}$, C^+ are closed and $C^+ \cap C^- = \{(0,0)\}$.

It is also possible to obtain similar results when $\pi(0,0) \neq 0$ and some $\mathcal{S}(M,\Omega)$ -varieties are available. For example we can prove the following theorem (cf. [32]).

Theorem 2.9. Let $\mathbb{R} \times \mathbb{R} \times \mathbb{R} \rightarrow \mathbb{R}$ compact and suppose there exists $\mathbf{R} > 0$ such that $u + T(\mathbf{R},u)$ implies $|u|_H < \mathbf{R}$, $\mathbb{R} \in \mathcal{S}(M,\Omega)$. If $\pi(\mathbf{R},\mathbf{R}) \neq 0$, then \mathbb{R} has two unbounded components in $\mathbb{R} \times \mathbb{R}$ and $\mathbb{R} \times \mathbb{R}$ bifurcating at $(0,\mathbf{R})$, where $\mathbb{R} \subset \mathbb{R}^+$.

2.3. GLOBAL BIFURCATION THEOREMS FOR POSITIVE SOLUTIONS.

It is FAIRLY well-known that positive solutions play a very important role in applications as, e.g., chemical reactions, combustion theory, population dynamics, etc., where most of the solutions (unstable limit cycles) are actually positive.

However, there are global bifurcation theorems for positive solutions ensuring the second possibility in Theorem 2.3, meaning in this way the existence of an unbounded component of (positive) solutions. Suppose $\mathbb{R} \times \mathbb{R}$ is ordered Banach space with positive cone \mathbb{R}_+ , P is generating ($0 \neq P \neq \mathbb{R}$) and let $\mathbb{R} \times \mathbb{R} \times \mathbb{R} \rightarrow \mathbb{R}$ compact such that $\pi(\mathbf{R},0) = 0$ for every $\mathbf{R} \geq 0$. We study the equation

$$(2.13) \quad u = \mathbf{R}u + T(\mathbf{R},u) \quad \text{on } \mathbb{R}_+$$

where L is a positive compact linear operator and \mathbf{R} is compact and such that $\pi(\mathbf{R},0) = 0$ ([63]). If $\mathbf{R} > 0$ uniformly for \mathbf{R} bounded,

Theorem 2.10 ([32]). Suppose all the above assumptions are satisfied and suppose that L generates strictly one positive eigenvalue λ_0 with a positive eigenvector. Then λ_0 is the only bifurcation point for positive solutions for (2.13). However, the set of positive solutions contains a unbounded component \mathbb{R} which is unbounded and such that $\mathbb{R} \subset \mathbb{R}^+ \times \{0\} \times \{1\} \times \{0\}$.

Remark 2.11. There are similar results concerning bifurcation at bifurcation points for positive solutions and take the corresponding conditions as in Theorem 2.8. There is a unique approach (infinity inf. [103]). In particular, it is possible to show that, under appropriate conditions to those in Theorem 2.8, there is a unique nonnegative solution in the case without bifurcation, say in case only one example.

Theorem 2.12 ([33]). Let $\mathbb{R} \times \mathbb{R}$ be an ordered Banach space with cone P and let $\mathbb{R} \times \mathbb{R} \times \mathbb{R} \rightarrow \mathbb{R}$ compact such that $\pi(\mathbf{R},0) = 0$ for every $\mathbf{R} \in P$. Then the connected component of the solution set of $u = \mathbf{R}u + T(\mathbf{R},u)$ containing $(0,0)$ is unbounded.

If, in addition, Theorem 2.7 becomes valid if $T(\mathbf{R},0) = 0$ for every $\mathbf{R} \in P$ in this case the variation we consider the "size" of trivial solutions $\mathbb{R} \times \{0\}$. It is interesting in this respect to study the existence of positive solutions bifurcating from the trivial solution.

An interesting nonlinear problem, where some of the above previous ones are valid is the equation

$$(2.14) \quad u = \mathbf{R}u + T(\mathbf{R},u) \quad \text{on } \mathbb{R}_+,$$

arising in the study of some simplified models for combustion (cf., e.g., [37]). This problem has been considered by several authors. In particular, Gouraud ([58]) proved that, in the special case of the sphere ($n=2$), there are indefinitely many solutions for a certain value \mathbf{R} of the para-

vector λ , and Zornberg-Joshiwara [8.3] provide a complete and interesting study of critical solutions by using ordinary differential equations methods.

Let $L = \{u\}^{\perp}$ and let γ be the semiball generated by $(\gamma, \mathcal{L}(L))$ associated to $\pi: L \rightarrow \mathbb{R}$ ($\pi(u) = u^\top u$). It is clear that (2.54) is equivalent to $u \in (\mathbb{M}_n(\mathbb{R}))^*$, or to $u = \pi(L)$, where $\pi(L) = \mathbb{M}_n(\mathbb{R})$ is compact, and that $\pi(L)u = 1$ for every $u \in L$. On the other hand, it follows immediately from the maximum principle that if u is a solution of (2.16) for $\lambda = \lambda_0$, then $u = 0$ in Ω . Indeed, again, since every solution is positive, the problem can be formulated in the above framework in particular, by theorem 3.7 it follows the existence of an unbounded continuum of positive solutions of (2.16) containing (0, 0).

This information, which is not so generic as we desire, can be completed by some additional considerations: first, there is $\lambda \in \mathbb{R} \setminus \{0\}$ such that (2.14) has no solution for $\lambda = \lambda_0$ (cf. Lemma 3.1 and [4.4]).

2.5. THE PRIMO POLYTOPES

It was already remarked that the Leray-Schauder degree is one of the most powerful tools to prove existence results for nonlinear problems, and that it can be useful to employ the same number of solvability theorems at least a lower bound. However, in some situations, e.g., in problems arising in physical applications, the main interest is concentrated about existence of positive solutions, and it would be nice to have at our disposal some analogous tools in order to work in a (relatively) open subset of a cone of positive functions. But, as will come out shortly (see, e.g., in $L^2(\Omega)$), this is not possible. The topological degree cannot be directly applied, since the corresponding open subset (the interior of the cone) is empty. Anyway, this difficulty can be overcome by exploiting the fact that a cone is a retract in a Banach space: this allows us to define a fixed point index for open operators defined on this cone.

First, we recall some definitions and thoughts from general topology. Let X be a topological space and let $A \subset X$; then A is called

a **subset** of X if there exists a continuous map $\pi: Y \longrightarrow X$ called **quotient** such that $\pi^{-1}(A) = A$. It is usually more than every retract is a closed subset.

Following a division by Deimling, every closed convex subset of a Banach space X is a retract of X . This allows us to define a **Floyd point index** (cf. [10.1.6b]) (for more details).

Let X be a retract of the Banach space E . Let U be a bounded open subset of X and let $f: \bar{U} \longrightarrow X$ compact which has no fixed points on ∂U . Then there exists an integer $I(f, U, X)$ with the properties:

1. **monotonicity:** For every inclusion map $\ell: \bar{U} \longrightarrow V$,

$$I(f, U, X) = 1.$$

2. **additivity:** For every pair of disjoint bounded open subsets U_1 and U_2 of X such that \bar{U} has no fixed points on $\bar{U} = U_1 \cup U_2$,

$$I(f, U, X) = I(f, U_1, X) + I(f, U_2, X)$$

where $I(f, U_1, X) = I(f|_{U_1}, U_1, X)$, $I = \pi_* \circ \ell$.

3. **homotopy invariance:** For every compact interval $I \subset \mathbb{R}$ and every compact map $R: I \times \bar{U} \longrightarrow X$ such that $R(I, t) \neq t$ for $(I, t) \in I \times \partial \bar{U}$,

$$\text{L}(R|_{I \times \bar{U}}, X) = \text{L}(f|_{\bar{U}}, X).$$

4. **dimension:** If Y is a retract of X and $\text{L}(Y, Y) \neq 1$, then for every $\lambda \in \mathbb{R}$,

$$\text{L}(E, \lambda Y, X) = \text{L}(E|_{\lambda Y}, Y, X)$$

5. **extension:** For $Y \subset U$, Y bounded and open, such that \bar{U} has no fixed points in $\bar{U} \setminus Y$,

$$\text{L}(E, \lambda Y, X) = \text{L}(E|_Y, Y, X)$$

where $\text{L}(E|_Y, Y, X) = \text{L}(E|_{\bar{U} \setminus Y}, Y, X)$.

5. **retraction:** If $\pi: U \longrightarrow Y$ is a fixed point index for open operators defined on U .

This integer is called the **Floyd point index**. In the case of an

measured Banach space $(\mathcal{B}, \|\cdot\|)$, the class $\mathcal{P} \subset \mathcal{B}$, following a theorem by Dini [10], a compact or \mathcal{B} , and then the corresponding closed pointwise limit will be denoted for a bounded open subset $\Omega \subset \mathcal{P}$ and a compact map $\mathbf{r}: \Omega \rightarrow \mathcal{P}$ which at fixed points in Ω . In the following the final point index will be denoted by $\text{lim}_{\mathcal{B}}$, omitting the reference to the index \mathcal{B} .

Remark 2.5: If \mathcal{B} is an open subset of the space \mathcal{D} , then $\text{lim}_{\mathcal{B}}(0, \mathbf{r}) = \text{d}(\mathbf{r}(0), 0)$. This shows that the time point index is a natural generalization of the degree. Cf. [10].

2.7. AN APPLICATION: POSITIVE SOLUTIONS FOR A REACTION-DIFFUSION EQUATION

This section is devoted to an application of the fixed point index in the selection of positive solutions of a nonlinear system arising in optimization, more precisely in the study of chemical catalysis (cf. [6, 12]). A variant of the standard global bifurcation theory is used for our problem, namely the following.

Lemma 2.7 ([8, 12]). Let $(\mathcal{B}, \|\cdot\|)$ be an ordered Banach space with cone \mathcal{P} and let $\mathcal{L}: \mathcal{B} \rightarrow \mathcal{B}$ compact, suppose that there is a constant $R > 0$ such that $\mathcal{L}w \geq R w$ implies $\|w\| \leq R$ and that $\{\mathcal{L}(0, \cdot)\}$, $\mathcal{L}(0) \times \{0\} \neq \emptyset$; then there exists a continuum of solutions of $\mathcal{L}w = R w$, which is unbounded in \mathcal{B}^+ if R and such that C contains $(0, 0)$, where $\mathcal{L} \in \mathcal{P}$ and $R \in \mathbb{R}_+$.

We shall apply this abstract result to our example. Consider the system

$$(2.15) \quad \begin{aligned} -u_1' &= u + \frac{u_1}{u_2^2} + \lambda u_1 = 0 & \text{in } \mathcal{B}^+ \\ (2.16) \quad -u_2' &= u_2 - 1 & \text{in } \mathcal{B}^+ \\ (2.17) \quad u &= v = 0 & \text{on } \partial \mathcal{B}^+ \end{aligned}$$

where \mathcal{B} is a convex bounded domain in \mathbb{R}^2 , $u_1, u_2, v > 0$ are real.

Moreover and λ is a real parameter. Let $\ell(w) = \frac{\|w\|}{w^2} + \lambda w$ it is clear that for $w \in \mathcal{B}$, $\ell(w) \geq \lambda$. We fix $\alpha_1 = \frac{1}{2}$ and λ and take \mathcal{B} as a parameter.

We consider the function space $\mathcal{B} = (\mathbb{C}^{1,0} \cap \mathbb{H})^2$ with the positive cone $\mathcal{B} = \{(w_1, w_2) \in \mathcal{B} : w_2 \geq 0\}$. We define a map

$$\mathcal{Q}: \mathbb{R}^d \times \mathcal{B} \longrightarrow \mathcal{B}$$

$$\mathcal{Q}(x, w) = (w_1, w_2 + Q(x, w))$$

In the following way, for each $(\lambda, w) \in \mathbb{R}^d \times \mathcal{B}$, $(w, 1)$ is the unique positive solution of the system

$$(2.18) \quad \begin{aligned} -w_1' &= w + \lambda w_1 = \lambda(w\ell(w)) & \text{in } \mathcal{B}^+ \\ (2.19) \quad -w_2' &= w_2 + \lambda w_2 = 1 & \text{in } \mathcal{B}^+ \end{aligned}$$

Indeed, for $w, v \in \mathcal{B}$, the system is decoupled in two linear equations. Moreover, by the Rabinowitz principle, $\mathcal{B} = \mathcal{B}^+ \cup \mathcal{B}^-$ (recall that $\mathcal{B}^+ \cup \mathcal{B}^- = \mathcal{B}$, $0 \in \mathcal{B}^-$). The operator \mathcal{L} is compact, and thus $\text{deg}_{\mathcal{B}} \mathcal{L}$ is easily straightforward.

Lemma 2.8. There is a solution $\theta \in \mathbb{R}$ which is independent of λ , such that if (λ, θ) is a solution of $(2.18) \times (2.19)$ with $\lambda > 0$ and $w, v \neq 0$, then

$$(2.21) \quad \left| \frac{\|w\|}{w^2} \right| \leq \sigma \quad \left| \frac{\|v\|}{v^2} \right| \leq \sigma \quad \forall \lambda \in \mathbb{R}.$$

Indeed, let q be the solution of the problem

$$(2.22) \quad \begin{aligned} -q' &= \lambda q - \sigma^{-1} & \text{in } \mathcal{B}^+ \\ (2.23) \quad q &= 0 & \text{on } \partial \mathcal{B}^+ \end{aligned}$$

By adding equations (2.21)–(2.23) and (2.22)–(2.23) we obtain

$$\delta(\eta_1 u + \eta_2 v + \eta) = -\lambda^2 \eta \text{ for } \eta \in \mathbb{R}.$$

$$\eta_1 u + \eta_2 v + \eta = 0 \quad \text{on } \partial\Omega,$$

Hence, by the maximum principle, $\eta_1 u + \eta_2 v + \eta < 0$ in Ω , and then $\eta_1 u + \eta_2 v + \eta < 0$, which gives the result.

Lemma 2.2. *There is a constant $R > 0$ such that if $(u, v) \in \mathbb{R}$ is a solution of (2.18)-(2.19) for $\lambda = \lambda_0$, then*

$$(2.24) \quad \|u\|_{L^2} \leq \frac{R}{2}, \quad \|v\|_{L^2} \leq \frac{R}{2}.$$

PROOF: It follows easily from (2.21), the L^p estimates in [1] and Harnack's lemma.

Theorem 2.3. *There exists an unbounded sequence of positive solutions of (2.18)-(2.19) such that its projection on the real axis is bounded by λ_0 . In particular, for any $\lambda > 0$ there is at least a positive solution of (2.18)-(2.19).*

PROOF: We apply Lemma 2.2 with $p = k$, $P = R$, and $\Gamma = \emptyset$. We know (by applying Lemma 2.1 with $p = k$) that $\|u\|_{L^k} = \|v\|_{L^k} = 0$ implies (2.24) (Lemma 2.2). We claim that $\lambda(u, v) = \lambda(u, v) = 0$. Indeed, recall that $\delta(u, v) \geq \lambda$ given by (2.18)-(2.19) with $\lambda = 0$ and derive the identity $\lambda(u, v) = \lambda(u, v)$ as the solution of the linear system

$$\begin{aligned} u + \lambda u &= 0 \quad \text{in } \Omega \\ -\Delta u + \lambda u &= 1 \quad \text{in } \Omega \\ u = v = 0 &\quad \text{on } \partial\Omega. \end{aligned}$$

Since $\lambda(u, v) = \lambda(u, v)$ and, moreover, (2.24) implies $\|u\|_{L^k} = \|v\|_{L^k} \neq 0$, for any $\lambda \in (0, \lambda_0)$ and any $(u, v) \in \mathcal{H}(0) \times \mathcal{H}_0$, the homotopy λ is a branch of the index yields

$$\lambda(\mathcal{G}(\lambda, 0), \mathcal{H}(0)) = \lambda(\mathcal{G}(\lambda, 1), \mathcal{H}_0(0)) = 1 \neq 0, \quad \text{so } \lambda_0 = 1.$$

Since $\lambda(u, v)$ is a constant w.r.t. the normalization frequency, finally, we remark that uniform bounds for any $\lambda > 0$ can be easily obtained.

Remark 2.3. We point out that in the example (2.18)-(2.19), the dissipativity of the reaction $\frac{\partial u}{\partial \nu} = u - v$ raises serious problems concerning the associated Neumann operator. This is no more the case if we are restricted to $u > 0$. On the other hand, the fact that we only consider positive solutions makes easier to obtain uniform bounds (cf. the proof of Lemma 2.3). Similar arguments, but without the whole space and not on a cone, were used in [42] for the Brusselator-

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continuation and variational methods are presented in their respective sections. Section 3.3 treats our test problem (the nonlinear eigenvalue problem (3.11)-(3.13)) by following a different approach involving the use of a local inversion theorem by CRANDALL-LABRKOVIC and a continuation argument which is an application of the NAGELIS function theorem.

Once again, numerical estimates play an important role. This method is particularly well-suited for the study of bifurcation from simple eigenvalues, as it is the case here. We also study some variants of problem (3.11)-(3.13), which led to some open problems.

Moreover, the same kind of analysis can still useful in presence of a small bifurcation. In Section 3.2 we consider a class of quasilinear equations (3.48). We only discuss the arguments leading to (48) for a careful study of the "bifurcation points". Finally, Section 3.3 introduces a new method, namely Lyapunov-Schmidt-like critical point theory, which, maybe combined with more other tools, is very useful to prove bifurcation and multiplicity results.

3.1. LOCAL IMMERSION THEOREMS AND CONVERGENCE: THE ALGEBRAIC CASE.

We will consider here, for the first time, the nonlinear eigenvalue problem (3.11)-(3.13) which was treated by using sub and superfunctions in Section 1.2 and by using degree theory and global bifurcation theories in Sections 2.2 and 2.3, respectively. We remark that, even if the idea of using local (local) and global immersion theorems to prove existence results is not very recent (BIRMAN, CACCIOPPOLA, etc.), we follow here the approach by NAKAMURA [90], cf. also [42].

Consider once again the problem

$$(3.1) \quad -\Delta u + f(u) = g \quad \text{in } \Omega,$$

where Ω is a smooth bounded domain in \mathbb{R}^n , $1 \leq n \leq 3$, f real parameter, and $g : \Omega \rightarrow \mathbb{R}$ satisfies

$$(3.2) \quad \begin{cases} \text{f is \mathcal{C}^1 increasing, and $f(0) = f'(0) = 0$}, \\ \text{$|f''(u)|$ is strictly increasing for $u > 0$ and strictly decreasing for $u < 0$}. \end{cases}$$

$$(3.3) \quad \begin{cases} \lim_{u \rightarrow +\infty} \frac{|f(u)|}{|u|} = +\infty \\ \lim_{u \rightarrow -\infty} \frac{|f(u)|}{|u|} = +\infty \end{cases}$$

or

$$(3.4) \quad \lim_{|u| \rightarrow +\infty} \frac{|f(u)|}{|u|} = +\infty \quad \text{and} \quad \lim_{|u| \rightarrow -\infty} \frac{|f(u)|}{|u|} = +\infty.$$

LEMMA 3.1. Assume that f satisfies (3.2), (3.3) and (3.4). Then (3.1), that is, if u is a nontrivial positive solution of (3.1), (3.2), (3.3), then $u_1 = \lambda_1 - \lambda_2 + \lambda_3 + \mathcal{O}(r^2)$.

PROOF. The first part is just Lemma 1.4. This would follow up a similar computation argument.

The following theorem is the main tool for the results in this section (cf. also [42], [18], [19]).

THEOREM 3.1 ([42]). Let X and T be real Banach spaces, let $F : T \times X \rightarrow \mathbb{R}$ a bounded bilinear and let $\Gamma : T \times X \rightarrow Y = T^\perp \subset \mathbb{C}^T$. Let $\beta \in \mathbb{C}$ and assume that F satisfies

- (1) $F(\beta, 0) = 0$ for every $\beta \in T$,
- (2) $\dim \ker F(\beta, 0) = \text{codim } \text{Ker } F(\beta, 0) = 1$,
- (3) $\text{dist} \ker F(\beta, 0) \cap \text{codim } \text{Ker } F(\beta, 0) > 0$,
- (4) $F(\beta, 0) \neq 0$ if $\text{Ker } F(\beta, 0) = 0$, where $\beta \in \mathbb{C} \setminus \{0\}$

(this will be denoted by $\{F_\beta\} = \text{Ker } F(\beta, 0)$).

Let \mathbb{D} be a compactly supported \mathcal{C}^1 -function in \mathbb{C} . Then there exist an interval I containing the origin and two \mathcal{C}^1 -functions $\lambda : I \rightarrow \mathbb{D}$ and $\gamma : I \rightarrow \mathbb{D}$ such that $\lambda(0) = \beta$, $\gamma(0) = 0$ and

$$\mathbb{P}_0(\lambda_1, \lambda_2) = \lambda_1 - \lambda_2$$

and $\pi(x) = \infty + \pi(y)$. Lemma [\(3.1\), \(3.5\)](#) implies $\pi^{-1}(0)$ is a neighborhood of $\pi^{-1}(0)$ by the curve $x = 0$ and $\{(x, \pi(x)) : x > 0\}$.

Theorem 3.2 ([41]). Suppose that \mathbf{f} satisfies [\(3.1\)](#), [\(3.4\)](#), [\(3.5\)](#) (resp. [\(3.8\)](#)). Then for any λ such that $\lambda_1 < \lambda$ (resp. $\lambda_1 < \lambda < \lambda_2$ if $\mathbf{f}'(y)$ above admits a unique nontrivial positive solution with $y > 0$) there exists a unique nontrivial positive solution with $y > 0$ from $\mathbf{f}^{\lambda_1, \lambda_2}$ (resp. from $\mathbf{f}^{\lambda_1, \lambda_2, \lambda}$). Moreover, if \mathbf{f} satisfies [\(3.1\), \[\\(3.5\\)\]\(#\) and \$\mathbf{f}'\(y\)\$ above admits a unique nontrivial positive solution with \$y > 0\$ then \$\mathbf{f}^{\lambda_1, \lambda_2, \lambda}\$ has a unique nontrivial positive solution with \$y > 0\$.](#)

(3.6). [\(3.6\)](#)

$$\begin{aligned} & \frac{d}{dx} (\ln |\pi(x)|)_x = u_x \\ & \frac{d}{dx} (\ln |\mathbf{f}'(x)|)_x = u_x \end{aligned}$$

PROOF. First, we recall that the uniqueness was already proved in [Theorem 3.1](#).

In the first part of the proof, the local monotonicity (Lemma [3.1](#)) is applied to show the existence of a single branch of positive solutions bifurcating to the right from $\lambda = \lambda_1$. However, this branch can be parameterized by λ .

We derive the implicit form

$$\mathbf{f}^{\lambda_1, \lambda_2}(x) = (x + \mathcal{C}^{2,3}(0))x = 0 \quad \text{on } \mathbb{R}.$$

and the mapping

$$\mathbf{f}^{\lambda_1, \lambda_2}(0) = (x + \mathcal{C}^{2,3}(0))x = 0 \quad \text{on } \mathbb{R}.$$

By

$$\mathbf{f}'(x) = -\Delta x + \mathbf{f}(x) = 0,$$

it is clear that $\mathbf{f}'(0) = 0$ for any λ and \mathbf{f}' is \mathcal{C}^2 by [\(3.31\)](#). It is easily calculated that

$$\mathbf{f}'_0(0, \pi(y))y = \Delta y + \mathbf{f}'(0)y = 0$$

and

$$\mathbf{f}'_0(0, \pi(y))y = y - w.$$

On the other hand, $\mathcal{E}_q(\lambda_1)$ is an involution. Indeed, it is not hard to show that $\mathcal{E}_q(\lambda_1) = \mathcal{E}_q(\lambda_2)$ with $\lambda_1 > \lambda_2$.

$$(13.6) \quad \begin{aligned} &= \frac{1}{2} \lambda + \mathcal{E}^2(\lambda) + \lambda \\ &\quad - \lambda = 0 \quad \text{on } \mathbb{R}. \end{aligned}$$

Thus, by (13.3), (13.4), $\mathcal{E}_q(\lambda_1) = \mathcal{E}_q(\lambda_2)$, and taking $\lambda > 0$, we obtain by usual compactness arguments

$$\lambda = \lambda_q(\mathcal{E}^2(\lambda)) = \lambda_q(\mathcal{E}(\lambda)) = \lambda.$$

which is a contradiction. Hence $\mathcal{E}_q(\lambda)$ is an involution and we apply the Involution Function Theorem to obtain the existence of $m < \nu > 0$ and $a \in \mathbb{C}^2$ helping $\lambda = m + \nu i$ defined for $|i| = a < c$ such that $\mathcal{E}_q(\lambda)(i) = 0$.

We have to prove that the solutions obtained prolongating in this way our initial branch are actually positive. By the continuity of $\lambda = \lambda_q(\cdot)$ it will be sufficient that if $\lambda = \lambda_q(\mathcal{E}^2(\lambda))$ then $\mathcal{E}^2(\lambda) = \lambda$. Let us do it for all the values of λ , then there should be a $\tilde{\lambda}$ maximal such that $\lambda_q < \tilde{\lambda}$ and $\mathcal{E}^2(\lambda) = \lambda$, i.e., $\mathcal{E}^2(\lambda)(\mathbf{x}) = 0$ for some $\mathbf{x} \in \mathbb{R}^n$. By Corollary 1.3, $\mathcal{E}^2(\lambda) \geq 0$ and moreover $\|\mathcal{E}^2(\lambda)\|_{\mathbb{R}^n} = 0$ when $\lambda = \tilde{\lambda}$. This implies (cf. Theorems 2.6 and 10.2) that $\tilde{\lambda}$ is the only bifurcation point for positive solutions and $\lambda = \lambda_q$, which is impossible.

Hence we have shown the existence of a branch of positive solutions for every λ in a maximal interval of the form $(\lambda_1^-, \lambda_1^+)$, where $\lambda_1^- < \lambda_1^+ < \lambda_2$. This implies that $\lambda = \lambda_q(\mathcal{E}^2(\lambda))$ is the only bifurcation point for positive solutions of our problem. By the way, one can proceed similarly for the branch of negative solutions. In particular, we improve substantially the results concerning the smoothness of the branch claimed in Corollary 1.3. It is also claimed in Proposition 1.1 that $\mathcal{E}_q(\lambda)$ is a function of λ in the class C^1 (and $\mathcal{E}^2(\lambda)$ is $C^{1,1}$).

Now we study the same problem but with \mathcal{E} strictly concave instead of strictly convex (assuming (13.4)), and try to apply the same methods.

More precisely, we assume that $\mathcal{E}: \mathbb{R} \rightarrow \mathbb{R}$ satisfies the same conditions, i.e. (c), there are a constant c and a sequence λ_n such that $\lambda_n \rightarrow \lambda$ and $\mathcal{E}(\lambda_n) \geq c$. A rather straightforward argument involving the comparison of the linear operator (or iteration operator) of (13)-(13.2) shows that it is possible to apply the Inversion Function

Theorem in \mathcal{X}^+ , maintaining the maximality of λ^+ . By (13.3) and the results in (6.1) it follows that $\lambda^+ > \lambda$ as supposed in iteration point, and then $\lambda = \lambda_q(\mathcal{E}^2(\lambda))$ (cf. 10.11, 11.12) for details.

At last, we mention (13.3), (13.4), (13.5). Then it is clear that a necessary and sufficient condition for $\lambda^+ > \lambda$ is the existence of a $\mathcal{E}^2(\lambda)$ -estimate for solutions of (13.1), (13.2), i.e., the existence of a continuous function $\# : \mathbb{R}^+ \rightarrow \mathbb{R}^+$ such that

$$(\mathcal{E}(\lambda))^2 \leq \#(\lambda).$$

The existence of such a $\#$ follows easily from Lemma 2.2 and the small regularity theory (cf. also Section 1.2). However, as we only want $\mathcal{E}^2(\lambda)$ -estimates for positive solutions, we may as alternative method, suppose that $\mathbf{w} \in \mathbb{R}$ is a solution for the value λ of the parameter. Let $R = \max_{\mathbf{x} \in \mathbb{R}^n} |\mathbf{w}(\mathbf{x})| > 0$. For some $\alpha > 0$, there

$$-\mathcal{E}(R\mathbf{x}) + \mathcal{E}(R\mathbf{x}) - \mathcal{E}(\mathbf{w}(\mathbf{x})) > 0.$$

By (13.4), $\mathcal{E}(R\mathbf{x}) = R\mathbf{x}$, and with the notation of Section 1.2, we have $R < g(R)$. Since $L = \mathcal{E}'(\mathbf{x})$ remains, together with the \mathcal{E}^2 and \mathcal{E}^3 -regularity estimates, global C^∞ (and even $C^{2,1}$) solutions.

The proof of Theorem 2.2 gives an alternative approach to the study of positive solutions of our problem. By the way, one can proceed similarly for the branch of negative solutions. In particular, we improve substantially the results concerning the smoothness of the branch claimed in Corollary 1.3. It is also claimed in Proposition 1.1 that $\mathcal{E}_q(\lambda)$ is a function of λ in the class C^1 (and $\mathcal{E}^2(\lambda)$ is $C^{1,1}$).

Now we study the same problem but with \mathcal{E} strictly concave instead of strictly convex (assuming (13.4)), and try to apply the same methods.

More precisely, we assume that $\mathcal{E}: \mathbb{R} \rightarrow \mathbb{R}$ satisfies the same conditions, i.e. (c), there are a constant c and a sequence λ_n such that $\lambda_n \rightarrow \lambda$ and $\mathcal{E}(\lambda_n) \geq c$. A rather straightforward argument involving the comparison of the linear operator (or iteration operator) of (13)-(13.2) shows that it is possible to apply the Inversion Function

$$\begin{aligned} & \text{LEMMA 3.12: } \lim_{n \rightarrow \infty} \|x_n\| = \ell^+(u) \\ & \quad \text{if } u \neq 0 \text{ or } u = 0 \text{ and } u \neq v. \end{aligned}$$

If u is nearly zero than it has zero

$$\begin{aligned} & \text{LEMMA 3.13: } \lim_{n \rightarrow \infty} \|x_n\| = \ell^+(v) \\ & \quad \text{if } v \neq 0 \text{ or } v = 0 \text{ and } v \neq u. \end{aligned}$$

A first result concerning non-existence of positive solutions can be proved by using the same contradiction arguments as in Lemma 3.4.

LEMMA 3.14: Assume that ℓ satisfies (3.10)-(3.12), $\min_{\Omega} \lambda^+ <$

λ^+ a non-trivial solution of (3.1), (3.2), $\lambda_1 + \ell^+(u) < \lambda^+ < \lambda_1 + \ell^+(v)$.

If ℓ is odd and monotonic differentiability retains in the study of the uniqueness of non-trivial positive solutions, which is essential to apply this continuation method, since none of the proofs of theorem 3.3 works.

Indeed, now the function $\ell = f^\pm$ is convex and odd and asymptotic line are no more available. But (local) Brouwer 1.5), the direct part of proof of Theorem 3.1 will still work if u and v are non-trivial positive solutions of (3.1)-(3.2) such that, e.g., $u \leq v$, then by (3.13) the same argument yields $u \leq v$. On the other hand, the second part, involving odd and even solutions, does not work. However, it is possible to prove that for two non-trivial positive solutions u and v , $0 < v < u$, then we have uniqueness. This will be a consequence of the following additional assumption

$$(3.15) \quad \lambda_1 + \ell^+(u) < \lambda_2 + \ell^+(v)$$

which can be written equivalently as $\ell^+(u) - \ell^+(v) < \frac{\lambda_2 - \lambda_1}{2}$. This assumption is exactly of the same kind of those in (3.4) or (3.7).

LEMMA 3.15: Assume that ℓ satisfies (3.10)-(3.12) and (3.15). If u and v are non-trivial positive solutions of (3.1), (3.2), then

$u < v$ or $v < u$.

PROOF: If u and v are solutions we have

$$\begin{aligned} & \Delta^2(u) + \ell(u) = \lambda_1 u + g_1 \\ & \Delta^2(v) + \ell(v) = \lambda_2 v + g_2 \end{aligned}$$

$$\begin{aligned} & \text{and if we put } u = 0 = v, \text{ this can be rewritten as} \\ & \Delta^2(u) + \ell(u) = \lambda_1 u \quad u = 0 \\ & \Delta^2(v) + \ell(v) = \lambda_2 v \quad v = 0 \end{aligned}$$

where x is defined as follows

$$x(\alpha) = \begin{cases} \frac{\ell(x(\alpha)) - \ell(x(\beta))}{\alpha(\alpha) - \beta(\alpha)} & \text{if } x(\alpha) \neq x(\beta) \\ x(\beta(\alpha)) & \text{if } x(\alpha) = x(\beta) \end{cases}$$

Then x is continuous and (3.15) and the definition of a map $\ell^+(\cdot)$ $\ell^+(\alpha) < x < \ell^+(\beta)$, $\alpha \neq \beta$.

We claim that $\ell^+\beta$ is invertible

$$\lambda_1 + \ell^+(u) < \lambda_2 + \ell^+(v)$$

then x is the fixed approximation of problem (3.13). Indeed, if we take α comparison results imply

$$\lambda_1 + \lambda_2(u) > \lambda_2(\ell^+(u)) = \lambda_2 + \ell^+(u)$$

so $\ell^+(u) < \ell^+(v)$. Hence the corresponding eigenfunctions $u = u - v$, 0 does not change sign in Ω and, $u = 0$ either $u > 0$, i.e., $u < v < 0$ either $w = 0$, $u = w$.

THEOREM 3.16 (3.12). Suppose that ℓ satisfies (3.10)-(3.12) and (3.15). Then for any λ such that

$$(3.16) \quad \lambda_1 + \ell^+(u) < \lambda < \lambda_2 + \ell^+(v)$$

there exists a unique non-trivial positive solution $u(\lambda)$ of (3.1), (3.2). The mapping $\lambda \mapsto u(\lambda)$ from $(\lambda_1 + \ell^+(u), \lambda_2 + \ell^+(v))$ into $C_0^2(\Omega)$ is C^1 . Moreover

$$(3.17) \quad \lim_{\lambda \rightarrow \lambda_1 + \ell^+(u)} \|u(\lambda)\|_0 = +\infty,$$

(iii) the uniqueness follows easily from Lemma 3.2 and 3.3. The existence proof is quite similar to that in theorem 3.2, so we only sketch it.

Firstly, by a completely similar application of theorem 3.1 we obtain the existence of a "small" branch of nontrivial positive solutions in a neighborhood of $\lambda_1 + \mathcal{E}'(0)$. This branch can be parametrized by λ' , and one can verify by the same reasoning again the implicit function theorem that $\varphi_{\lambda'}(u, v)$ is a solution with $v > 0$, where (u, v) is a solution with $v < 0$ at time $\mathcal{E}'(0) = \frac{\pi}{2}L^2$ by (3.13). The proofs that the solutions on this branch satisfy all the properties that there is a unique λ'' for this continuation, and that $\lambda'' = \lambda_1 + \mathcal{E}'(v)$ are analogous.

The existence results in theorems 2.2 and 2.3 can be summarized in the following diagram (cf. Figures 3 to 5). We point out that $\lambda_1 + \mathcal{E}'(v)$ is an increasing function of v . In the first case (cf. Proposition 1.7) but we do not know if this is true, i.e., if $\lambda(v)$ is decreasing in v , under the assumptions of theorem 3.3, in Fig. 3.

Figure 2 corresponds to theorem 3.2 with assumption (D-5), and Figure 4 to the same theorem with assumption (D-6). Figure 5 illustrates the situation for theorem 3.3.

To estimate the value of a function \mathcal{E} satisfying the same hypothesis as (3.13)-(3.15), but not the additional hypothesis (3.14), which play a decisive role in the uniqueness proof. Suppose that \mathcal{E} satisfies $(3.14) - (3.15)$ and (3.16).

$$\lambda_2 + \mathcal{E}'(v) < \lambda_1 + \mathcal{E}'(0).$$

In this situation the first part of the argument is identical, and we prove exactly as before the existence of a "small" branch of nontrivial positive solutions, parametrized by λ' , in a neighborhood of $\lambda_1 + \mathcal{E}'(0)$. However, if $\varphi_{\lambda'}(u, v)$ is a solution with $v > 0$ and $\varphi_{\lambda'}(u, 0) + \mathcal{E}'(0) = 0$, then $\varphi_{\lambda'}(u, v)$ is an isomorphism (the proof is the same) and one implicit function theorem can be applied. In this way it can be proved, by using the fact $\lambda_2 + \mathcal{E}'(v)$ is an asymptotic bifurcation

point for positive solutions, that for any v in the interval $(\lambda_1 + \mathcal{E}'(0), \lambda_2 + \mathcal{E}'(0))$ there is a unique nontrivial positive solution. However, we do not know how to estimate this branch with the "small" branch of positive solutions starting from $\lambda_1 + \mathcal{E}'(0) + \mathcal{E}(1)$. Figure 6. On the other hand, it was proved in [61] by using the routine in [53], that there is at least one continuous positive solution for any v in the interval $(\lambda_1 + \mathcal{E}'(0), \lambda_1 + \mathcal{E}'(1))$.

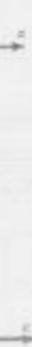


FIG. 3



FIG. 4

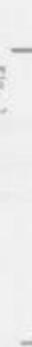


FIG. 5

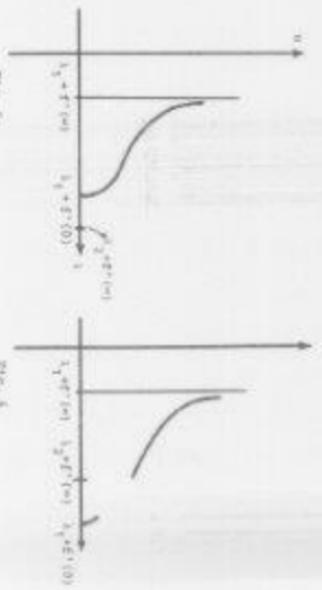


FIG. 2

BIFURCATIONS

We consider the nonlinear eigenvalue problem

$$(3.18) \quad -\Delta u + V(x, u) = \lambda u \quad \text{in } \Omega,$$

$$\text{with } u = 0 \quad \text{on } \partial\Omega.$$

where the function $V : \bar{\Omega} \times \mathbb{R} \rightarrow \mathbb{R}$ satisfies

$$(3.19) \quad F := \frac{d}{du} V(0, u) > 0 \quad \text{for any } u \in \mathbb{R},$$

$$(3.20) \quad E := C^1_c(\bar{\Omega}) \text{ for any } \lambda \in \mathbb{R},$$

$$(3.21) \quad E_{xx}(0, u) > 0 \quad \text{for any } u \in \mathbb{R} \text{ and any } x \in \Omega.$$

This kind of problems arises in several physical applications (cf. [393, 373]). An interesting particular case, namely $E(u) = u^{-\alpha}$, $\alpha > 0$, is considered in Sect. 2.5. For this purpose the map $E(u) = u^{-\alpha}$ ($1 + u^\alpha \neq 0$) is more difficult to treat (cf. [373], [443]).

It was pointed out in Section 2.5 that, for $\lambda > 0$, every solution u is strictly positive on Ω and that there are no bounded components of positive solutions emanating from 0. Moreover, by using well-known perturbations and the obvious fact that u is a sub-solution for (3.18), it follows the uniqueness of a minimal positive solution for some range of values of the parameter λ (cf. [393], [443]), but we are interested in getting some more information about solutions of (3.18)-(3.19).

Let $u_0 \times 0$ be the first eigenvalues of the linearized problem

$$-\Delta u + V_x(x, 0)u = \lambda u \quad \text{in } \Omega,$$

$$u = 0 \quad \text{on } \partial\Omega.$$

Then the mapping $\lambda \mapsto u(\lambda)$ from $(0, \tilde{\lambda})$ into $C^1_c(\bar{\Omega})$ is continuous with respect to the existence of a curve $\lambda \mapsto u(\lambda)$ of positive solutions with the properties

- (i) the mapping $\lambda \mapsto u(\lambda)$ from $(0, \tilde{\lambda})$ into $C^1_c(\bar{\Omega})$ is invertible as a homeomorphism;
- (ii) the operator $v \mapsto Av - \lambda E_{xx}(0, u)v$ is invertible as a Fredholm operator $\mathcal{L}_0^{1,0}(\bar{\Omega})$ into $\mathcal{L}_0^{1,0}(\bar{\Omega})$ for all $\lambda \in \mathbb{R}$.

It is interesting to see the behaviour of the curve of solutions near the critical value $\tilde{\lambda}$ of the parameter. The special cases $\alpha = 2$ and $(1 + u)^\beta$ were treated in [373] for spaces of weakly plane functions and some of these results were extended to general bounded domains in [443]. This is the main theorem in [443]:

Theorem 3.2. Suppose that $\alpha < 2$ such that $|u_0(\lambda)| \leq \lambda R$, where $|u_0(\lambda)|$ is the curve of minimal positive solutions given by Theorem 2.4. Then there exists $\tilde{\lambda} = \tilde{\lambda}(\alpha, R)$. In the topology of $C^1_c(\bar{\Omega})$, there exists $\tilde{\lambda} > 0$ such that the solutions of (3.18)-(3.19) near $(\tilde{\lambda}, 0)$ are on a curve $\lambda u(\lambda)$, $|\lambda| < \tilde{\lambda}$, with u and λ satisfying

(i) the map $\theta \mapsto (\lambda(\theta), \tilde{u}(\theta))$ from $(-\tilde{\lambda}, \tilde{\lambda})$ into $\mathbb{R} \times C^1_c(\bar{\Omega})$ is C^1 ,

(ii) $\lambda(\theta) \rightarrow \tilde{\lambda}$, $\tilde{u}(\theta) \rightarrow \tilde{u}$, $\tilde{u}'(\theta) \rightarrow v_*$, where v_* satisfies

$$-\theta \tilde{u}'' = \tilde{u}'(v_*)v_*, \quad \theta = 0 \text{ iff } \tilde{u}'(0) = 0, \quad \theta > 0 \text{ iff } \tilde{u}'(0) < 0.$$

It is clear that this theorem implies the existence of exactly two solutions in a deleted left neighbourhood of the critical value $\tilde{\lambda}$. The proof can be divided in two parts. Suppose that the curve of minimal positive solutions given by Theorem 3.2 has a limit point of the form $(\tilde{\lambda}, 0)$ ($\tilde{\lambda} < \tilde{\lambda}$). In the topology of $\mathcal{L}_0^{1,0}(\bar{\Omega})$, then by local coverings arguments as those in Theorem 3.1 it follows that $\lim_{\lambda \downarrow \tilde{\lambda}} u(\lambda) = 0$ and that the curve $u(\lambda)$ can be continued "smoothly" through $(\tilde{\lambda}, 0)$ ("bending back" precisely in (3.6)). In such a way that there are exactly

two other continuous and compactification arguments it is not difficult to prove the following result (cf. [393], [443]):

Theorem 3.3. If u is a solution of (3.18)-(3.19) with $\lambda > 0$,

two positive solutions in \mathbf{x} (obtained after rescaling) of \tilde{Y}_T , i.e. it is possible to prove that, under assumptions (3.20)-(3.22), $\tilde{Y}(x)$ is equivalent to the existence of a BESSEL estimate for the solution $u(x)$ to $u = \tilde{x} - \tilde{z}$. This is due very nearly that $\tilde{g}(x,t) = \tilde{g}_0^*$ and even more difficult for $\tilde{f}(x,t) = (\tilde{x} + \tilde{y})^{\frac{1}{2}}$ (cf. (4.4)).

3.3. EXISTENCE OF A SECOND POSITION: VARIATIONAL METHODS.

We aim in this section to improve the results of the preceding. For that, we make some supplementary assumptions which allow the use of a different kind of methods. Under those assumptions we shall prove that (3.18)-(3.19) has at least two solutions for any x in $(0,\tilde{x}_1)$. Assume that \tilde{x} satisfies (in addition,

$$(3.43) \quad \begin{aligned} & \tilde{f}(x,t) < c_1 + \frac{c_2}{2} t^{1/2} \quad \text{for } x \in [0, \tilde{x}], \quad \text{if } p \leq 2, \\ & \tilde{f}(x,t) < c_1^* \quad \text{with } \lim_{t \rightarrow \infty} \frac{\tilde{f}(x,t)}{t^{1/2}} = 0 \quad \text{if } p > 2, \end{aligned}$$

$$(3.44) \quad \begin{aligned} & \tilde{f}(x,t) = \frac{c_1}{2} t^{1/2} \quad \text{for } x \in [0, \tilde{x}], \\ & \tilde{f}(x,t) < c_1^* \quad \text{with } \lim_{t \rightarrow \infty} \frac{\tilde{f}(x,t)}{t^{1/2}} = 0 \quad \text{if } p > 2. \end{aligned}$$

$$(3.45) \quad \text{There exist } 0 < q < \frac{1}{2} \text{ and } \tilde{z} > 0 \text{ such that for } x \in$$

$$\tilde{g}(x,t) = \int_0^q \tilde{f}(x,u) du + \tilde{z} x f(u,x) -$$

Theorem 3.8 ([44]), if \tilde{x} satisfies (3.20)-(3.25), then (3.46)-(3.48) have at least two positive solutions for every t in $(0,\tilde{t})$.

Theorem 3.8 follows from a very well-known abstract result, the "Mountain Pass Lemma", due to Ambrosetti-Rabinowitz ([5]; cf. also [6]).

¹Problem 3.7 ("Mountain Pass Lemma"). Let E be a real Banach space and let $I \subset C([0,R])$ with $I(0) = 0$. Suppose that I satisfies

property (a) $\exists \rho, \delta > 0$ such that $I(u) > 0$ in $B_\rho(0) \cap E$ and $I(u) \geq \delta$ in $B_\rho(0)^c$.

(i) There exists $u \in E \setminus \{0\}$ such that $I(u) = 0$.

(ii) Palais-Smale condition: if $\{u_n\}$ is a sequence in E such that $I'(u_n) \rightarrow 0$, $I(u_n)$ is bounded and $I(u_n) \rightarrow I_*$, then u_n converges a convergent subsequence.

Then I has a critical point $\tilde{u} \in E$ such that $I(\tilde{u}) \geq I_*$.

²Proof of Theorem 3.8 ([44]). We redefine \tilde{E} for $x < \tilde{x}$ in such a way that $\tilde{x} = 0$ and $\tilde{g}(x,-)$ satisfies (3.23)-(3.25). Then the function $\tilde{v}(x) = \frac{1}{2} \int_0^x |\tilde{g}(u,-)|^2 du$ is the minimum of \tilde{E} and (3.24).

$$\tilde{v}(x) = \frac{1}{2} \int_0^x |\tilde{g}(u,-)|^2 du = \frac{1}{2} \int_0^x \tilde{f}(u,-) du$$

is $C_0^1([0,1]) \neq 0$ by the minimality of \tilde{x} and (3.24). We define another functional

$$J(u) = \frac{1}{2} \int_0^1 |u'|^2 dx - \int_0^1 \tilde{f}(u,-) dx$$

(for $0 < x < \tilde{x}$) exactly where $u(1)$ is the initial positive solutions. Then $J'(u) = 0$ and 0 is the critical point corresponding to \tilde{x} .

The relevance of a second solution follows from the application of Theorem 3.7 to the functional J : it is clear that $(\tilde{x}_1^*, \tilde{x}_2^*)$ will follow from assumptions (3.46)-(3.25).

CONTENTS AND BIBLIOGRAPHICAL NOTES

For a general overview of nonlinear diffusion equations we refer to the recent book by Souplet (2006), and also to the book (Barbu and Roubíček (2007)). A classical reference for bifurcation theory is the book by Krasnosel'skiĭ (1971). The recent book by Cîrstea-Schäfle (2007) covers many related topics, especially local bifurcation methods. Other similar treatments are the books by Berber (1972), Deimling (1974) and Protter-Morrey (1984), the lecture notes by Gómez (1981), Amann (1981) and de Giorgi (1981). Some many more references can be found.

With respect to α , the topological methods mentioned here are applicable to a very large variety of nonlinear problems. Among them we want to mention abstract interesting papers, namely Ambrosetti-Prodi type problems (cf., the survey by Ambrosetti (1981), and also (Ambrosetti 1982), (Ambrosetti 1983), (Ambrosetti-García-Lpez 1987)), and superlinear problems (cf., the book by Cotterell-Hastings (1971)). However, the method of sub and super solutions was only developed in the end of the sixties with papers by Olszak (1967), Codd-Godwin (1971), Peletier-Gmehl (1971), Sengpiel-Codd (1971), see (cf., the bibliography of (10)). A more general and systematic presentation was given by Amann (1971) and Bartsch (1974). An abstract version of the method in the framework of ordered Banach spaces was given some years after by Amann (1976).

As it was pointed out above, we do not try to offer a very general version of this method, but only a simplified one in order to show clearly how it works. The measure α can be replaced by more general second order linear differential operators with sufficiently smooth coefficients satisfying the maximum principle (cf., (13)(14)). The assumption that the nonlinear term F is C^2 is too restrictive, too. Instead we will work in L^2 or locally C^2 ($n = n + 1$) and there exists methods which work for only $n \geq 1$ (cf., a no. 10 increasing in (A7), (A8), (A9) (1970)). The method is also applicable to anisotropic conditions (1970) (even including unilateral constraints (63)) - 60

nonlinearity depending on the gradient (cf., (14) and its references) and to problems with discontinuous linear terms (cf., e.g., the work by Chang (1981) and Rabinowitz (1997)). There are many other approaches (integral and integro-differential equations, variational inequalities, etc.) for all which concerns our main tools, the Mountain principle, cf., the books by Zeidler-Blumke-Blumke (1981) and Gilbarg-Trudinger (1983).

Some interesting applications of this method to nonlinear elliptic problems can be found in the paper by Kastan-Moser (1991) (cf., also (48)). In this application were also applied in (48) to obtain δ -shaped bifurcation curves and in (10) for the study of perturbed bifurcation problems. More references in this book by Coddington-Hillert (1971) (cf., also (1961)) concerning the properties of the eigenvalues and the eigenfunctions of (1), (16)(17), maximum and minimum properties on the density and the co-codensity. In addition, many of the results in (9)-(10) or (16) are also applicable to this problem. A basic tool are the numerical results contained in the book by Coddington-Hillert (1971) (cf., also (1961)) concerning the properties of the eigenvalues and the eigenfunctions of

(1), (16)(17), maximum and minimum properties on the density and the co-codensity. In addition, the results in this section are taken from (61) (20), cf., also (192)(1967). The first uniqueness proof is a slight improvement of (17) and (18) (cf., the generalization in (61)). The second is included in (10). However, a full uniqueness proof, using Schrödinger's Brunnian principle, can be found in (192), (cf., (27) for a related result). The case of asymptotically linear F is treated in (61). The more multiplicity result of proposition 5.3 we find proved in (33) by reasoning as in section 3.2 before.

The results in Section 5.3 are contained in (62). A first solvability theorem for systems by using sub and superfunctions was already given by Bartsch (1981), cf., (1982)(1984)(1984), and the extension thereof for related results. Concerning the Mountain principle for systems cf., (1982)(62). An extension of the existence theorem in Section 5.3 was given in (53); this generalization was motivated by an study of some free boundary problems for reaction-diffusion systems, cf., (16)(19)(1970)(1981). Cf., also (45)(1971) for related existence results. Techniques for systems is a rather difficult question; some more of these partial results can be found in, e.g., (2)(1971)(2)(1971)(1971) and (the

process at the end of time $\Omega(1/\epsilon)$. Stationary solutions generate (1.125), or by global bifurcation arguments (cf., Section 3.2 and (3.2)). We do not include here the application of these methods, namely invariant relations, to the associated periodic problems, cf., the book by Sotomayor [105] and the work by Bautista [105] (1991), Astolfi [106], Chicone-Conte-Ortega [107], Astolfi et al. [107] (1992). The 0D approximation is a system arising in combustion theory cf., [193] (1961).

Concerning the stability of solutions of the stationary problem, it was proved by Hartinger [193] (1971) (cf., also (1.01)) in the case of a simple equation that if the solution obtained by sub and superpositions is unique, then it is stable and all solutions with initial data in the interval $[0, \infty)$ converge to this unique solution (one can refer to Hartinger [193] for more details). If there are multiple solutions, then the problem becomes interesting. If there are multiple solutions, then the problem becomes more involved (cf., e.g., [107]). For analytic results for systems with sub and superpositions etc., see [193] (1961).

A very nice presentation of topological degree theory is given in the lecture notes by Paluszak [97], where many applications are also included. A detailed and systematic treatment of the degree can also be found in the books by Dugundji [142] and Nemytskii [94] (cf., also [73], [177] (1970)). The results of Section 2.2 are due to LaSalle-Yoshizawa [76], cf. also [53]. For other multiplicity results see, e.g. [105] (1991) (4.8) and corresponding bibliographies.

The global bifurcation theorems in Section 2.3 are due to Zabrejko (cf., [17]-[93] (1971)), whose applications to nonlinear Steklov-Liouville problems and quasilinear elliptic equations are also included. Cf., also [316] (1971). The theorem is due to Zabrejko (cf., also [105] (1991)). For bifurcation at infinity, cf., [95]. The theorem is due to Krasnosel'skiĭ [191] (1972). Global bifurcation theorems are equally due to Bautista [105] (1991). Global bifurcation theorems for positive solutions were obtained independently by numer [147] and Turner [118], cf. also [105] (1991).

New information on the fixed point index is given in [20] and [205]. In particular, [20] contains a number of applications, cf., also [21] (where some results by Krasnosel'skiĭ [191] are extended) and [61].

Section 2.7 is a simplified version of [62], where the case of

nonlinear boundary conditions was treated. A variant of a predator-prey system considered by Cressman-Gilmer [197] is also studied in [61]. A 3-D, n -D was already pointed out in Remark 3.3 in the introduction to applying the fixed point index, mainly because in this case it is sufficient to obtain \mathbb{R}^n -valued estimates (cf., for example, points 3).

For other applications of the degree cf., [105] (1970) (4.8) [107]. The technique in Section 2.3 was used by Paluszak, cf., [97] (and [193] (1971)). It is possible to combine the method of sub and superpositions with the degree (for the fixed point index) to obtain more similar results in the literature [142] (1971) (4.8) [105] (1991). For inversion theorems for mappings with submersion and supermersion cf., [133] (1964) [182]. The theorem in Section 3.3 is due to [61], cf., [4.7] (1970) (4.8) [105] (1971). The main tool (Theorem 3.3) is contained in [4.7], but there are similar results in the literature [142] (1971) (4.8) [105] (1991). For inversion theorems for mappings with submersion and supermersion cf., [133] (1964) [182]. The theorem in Section 3.3 is due to [61], cf., [4.7] (1970) (4.8) [105] (1971). The main tool (Theorem 3.3) is contained in [4.7], but there are similar results in the literature [142] (1971) (4.8) [105] (1991). For inversion theorems for mappings with submersion and supermersion cf., [133] (1964) [182]. The theorem in Section 3.3 is due to [61], cf., [4.7] (1970) (4.8) [105] (1971). The main tool (Theorem 3.3) is contained in [4.7], but there are similar results in the literature [142] (1971) (4.8) [105] (1991). For inversion theorems for mappings with submersion and supermersion cf., [133] (1964) [182].

Bifurcations 3.2 and 3.3 contain results by Górnall-Paluszak [64], [65] (1971) (4.8) (for sections 3.2). These problems were treated by Dugundji [142] and Daugherty-Jordan [63] in the case of a single, often trival solution. For the "bifurcation" case see in [105] (1991). Similar problems have been studied by Bautista-Sánchez [67], cf., also [105] (1991) (28). The same results can be obtained by using plane phase portraits.

Bifurcations 3.2 and 3.3 contain results by Górnall-Paluszak [64], [65] (1971) (4.8) (for sections 3.2). These problems were treated by Dugundji [142] and Daugherty-Jordan [63] in the case of a single, often trival solution. For the "bifurcation" case see in [105] (1991). Similar problems have been studied by Bautista-Sánchez [67], cf., also [105] (1991) (28).

A good reference for the critical point theory used in Section 3.7 is the survey [96] by Paluszak, which contains some applications. The Mountain Pass Lemma is due to Rabinowitz-Paluszak [151], cf., also [141]. The other applications involving critical point theory with the previous methods (sub and superpositions, degree, etc., cf., [61], cf., [148] (1971)).

Our basic reason to consider semilinear eigenvalue problems of the

plied in this situation (cf., §5), improving some of the previous results.

Another interesting example is the predator-prey system

$$\begin{aligned} u_t &= \alpha u + \beta uv - \gamma u^2 \quad \text{in } \Omega, \\ v_t &= \delta v - \epsilon v u - \eta v^2 \quad \text{in } \Omega, \end{aligned}$$

instead of nonlinear problems of the type

$$\begin{aligned} u_t &= \lambda(u)u \quad \text{in } \Omega, \\ u &= 0 \quad \text{on } \partial\Omega, \end{aligned}$$

which are more frequently studied in the literature, relies on the fact that the problem of existence of positive solutions for some considerably different systems with homogeneous Dirichlet boundary conditions can be reduced to a similar perturbation of the former problem. (Indeed it is very clear that problems are not independent, and many of the results ref. n. 19) apply to both of them.)

More precisely, a system arising in the activator-inhibitor interaction model can be reduced by a desingularizing technique (cf. 1981[78, 103]) to the equation

$$u_t = \alpha u + \beta uv + \gamma u^2 - \delta u \quad \text{in } \Omega,$$

$$v_t = \delta v - \epsilon v u - \eta v^2 \quad \text{in } \Omega,$$

where $\alpha, \beta, \gamma, \delta$ are real parameters and ϵ and η are as above.

A slight variant of this system was studied in [40], and an open problem was addressed: all uniqueness of noncritical positive solutions for some range of the parameters. Results were given by Blasius [21] by using global bifurcation and by using (47), this time by using degree for cones. Here again continuation methods can be applied [45]. Related problems arising to the study of the spread of a bacterial infection (cf. 1981[22]) were treated by Blasius (cf. [24]; 25)), again by global bifurcation methods and sub and superfunctions.

where X is the adjoint operator of a linear elliptic equation and \mathcal{A} satisfies the weak (A_0) assumption. This problem was treated by Boile 1981 for odd \mathcal{A} by using critical point theory. At the same time [26] (1981) used degree theory for the same problem with \mathcal{A} not necessarily odd showing the results in section 2.2 are the specialization of those in (18) to $\mathcal{A} = 0$.

The principal difficulty to apply the method of sub and superfunctions to this problem is the existence of some kind of maximum principle for the nonlinear linear operator $\mathcal{A} - \delta + \mathcal{B}$. A review of this type together with applications to some nonlinear equations (including the predator-prey one given by de Vries-Goldberg-Metzger 1977). The same paper contains also an application of the minmax-Pear-Lions to related problems (cf. 1981[72]). For other results in the same direction, the local minmax and mountain pass methods in section 3.1 can also be applied in this situation (cf., §5).

REFERENCES

- [1] S. Agmon, A. Douglis and D. Nirenberg, Estimates near the boundary for solutions of elliptic partial differential equations satisfying general boundary conditions. I, *Comm. Pure Appl. Math.*, 12 (1959), 623-727.
- [2] H. Amann, On the existence of positive solutions of nonlinear elliptic boundary value problems. *Indiana Univ. Math. J.*, 21 (1972), 125-155.
- [3] H. Amann, Fixed points of asymptotically linear maps in ordered Banach spaces. *J. Funct. Anal.*, 14 (1973), 142-171.
- [4] H. Amann, Multiple positive fixed points of asymptotically linear maps. *J. Funct. Anal.*, 17 (1974), 476-523.
- [5] H. Amann, Semilinear elliptic equations with nonlinear boundary conditions. In: *Nonlinear analysis, function spaces and differential equations*. W. Bebernes (ed.), Amsterdam, North-Holland, 1976.
- [6] H. Amann, Nonlinear operators in ordered Banach spaces and some applications to nonlinear boundary value problems. In: *Handbook of Mathematics* (Ed. R. K. Jain), 1976, 1-55.
- [7] H. Amann, Semilinear eigenvalue problems having previously two solutions. *Trans. Amer. Math. Soc.*, 250 (1979), 225-237.
- [8] H. Amann, Elliptic eigenvalue problems and nonlinear eigenvalue problems via nonlinear boundary value problems. *J. Math. Anal. Appl.*, 65 (1979), 432-462.
- [9] H. Amann, Existence and stability of solutions to semilinear parabolic systems, and applications to some reaction-diffusion equations. *Proc. Roy. Soc. Edinburgh Sect. A*, 88 (1981), 35-47.
- [10] S. Asmussen, Fixed point equations and nonlinear eigenvalue problems in ordered Banach spaces. *Stoch. Proc. Relat. Areas*, 16 (1979), 620-639.
- [11] S. Asmussen and M. Cranford, On some existence theorems for semilinear elliptic equations. *Indiana Univ. Math. J.*, 27 (1978), 759-768.
- [12] S. Asmussen and T. Jägerle, Positive solutions of certain nonlinear elliptic problems. *Indiana Univ. Math. J.*, 25 (1976), 259-279.
- [13] A. Friedman and G. Kanso, Sharp minimum mass results for some semilinear problems. *Monatsh. Math.*, 77 (1978), 433-453.
- [14] A. Friedman and G. Kanso, On the interaction of some different types of singularities with applications between various spaces. *Forum Math.*, 26 (1973), 237-242.
- [15] A. Friedman and P.-M. Nohel, Dual variational methods in optimal point theory and applications. *J. Funct. Anal.*, 14 (1973), 269-311.
- [16] C. Gallo, R. P. Gilbert and Z. Shabtay, Diffusion and reaction with nonlinear kinetics. *Nonlinear Anal.*, 8 (1984), 321-333.
- [17] J. Guckenheimer, A mathematical analysis of some problems from combustion for parabolic systems. *Indiana Univ. Math. J.*, 28 (1979), 261-277.
- [18] J. Guckenheimer, K.N. Chueh and W. Pitsch, Some applications of interval arithmetic for periodic systems. *Indiana Univ. Math. J.*, 28 (1979), 279-295.
- [19] J. Guckenheimer and K. Kammeyer, Invariant sets and the Hopf-Kaup-Furter property for systems of parabolic partial differential equations. *Rocky Mountain J. Math.*, 7 (1977), 553-567.
- [20] R. Jerrard, On the number of solutions in traveling wavefronts and kink-like elliptic problems. *J. Funct. Anal.*, 43 (1981), 1-27.
- [21] R. Jerrard, R. Kusner and B. Simon, Traveling wave solutions to reaction-diffusion and competition systems. *Rocky Mountain J. Math.*, 12 (1982), 115-138.
- [22] R.E. Showalter, *Nonlinear Analysis and Convexity Analysis*, New York, Academic Press, 1984.
- [23] J. Slat and G.L. Brown, Influence of steady-state values on modeling production and competition systems. *Proc. Roy. Soc. of Edinburgh Sect. A*, 74 (1984), 21-34.
- [24] J. Slat and G.L. Brown, A start-up-decay system modelling the spread of temperate vegetation. To appear in *Math. Biosci.*, in Appl. Biol.
- [25] J. Slat and G.L. Brown, To appear.
- [26] C.M. Brauner and B. Nkonga, Self dual problems and related singular perturbation problems: one problem à trois faces. *Arch. Ration. Mech. Anal.*, 86 (1985), 121-171.
- [27] J. Slat and L. Oswald, Remarks on nonlinear elliptic equations. To appear.
- [28] K.-U. Böhme, Sparsität (heterogenes) analytische Funktionen für Systeme von elliptischen Differenzial-Gleichungen. *J. Math. Anal. Appl.*, 55 (1976), 251-264.
- [29] K.-U. Böhme, W.H. Brézis and R. Kusner, L-shaped heteroclinic curves. *Bollettino U.M.I.*, 9 (1984), 475-488.
- [30] K.-U. Böhme and B. Rüdiger, Simple proofs of some results in perturbed bifurcation theory. *Ricerche di Matematica*, 32A (1982), 71-82.
- [31] V. Capasso and L. Medaglia, Convergence to equilibrium states for a reaction-diffusion system modelling the spatial spread of a disease. *Indiana Univ. Math. J.*, 30 (1981), 513-526.
- [32] V. Capasso and L. Medaglia, Stable joint behavior for a reaction-diffusion system: application to a class of epidemic models. *Math. Nachr.*, 100 (1982), 580-597.

- (33) Hing-Cheung Chan. On the multiple solutions of the elliptic differential equations with discontinuous nonlinear terms. *Indiana Univ. Math. J.*, 31(1982), 139-156.
- (34) S.-H. Chen and J.-S. Gao. *Method of Substitution*. Dover, New York, Springer, 1981.
- (35) E.S. Coddington and J. Nohel. *Positive linear operators*. In *Topics in the theory of nonlinear diffusion equations*. Indiana Univ. Math. J., 27(1978), 117-132.
- (36) S.-H. Chen. Positive solutions of nonlinear eigenvalue problems: applications to nonlinear elliptic dynamics. *Arch. Rat. Mech. Anal.*, 78(1981), 355-375.
- (37) S.-H. Chen and T. Lantcher. Nonlinear boundary value problems suggested by chemical reactor theory. *J. Diff. Eqns.*, 7(1970), 217-226.
- (38) G. Cramer. *Explicit formulas and the Morse Index*. Cont. Report Math. Ser., 10, Amer. Math. Soc., Providence, 1976.
- (39) E. Conway and D. Sussmann. Diffusion and the predator-prey interspecies competition. *Arch. Math.*, 33(1979), 672-686.
- (40) E. Conway, P. Smereka and J. Smoller. Instability and bifurcation of steady-state solutions for predator-prey equations. *Arch. Rat. Mech. Anal.*, 79(1980), 289-326.
- (41) R. Courant and D. Hilbert. *Methods of Mathematical Physics*. New York, Interscience, 1953.
- (42) M. Crandall and P. Rabinowitz. A bifurcation from simple eigenvalues. *Arch. Rat. Mech. Anal.*, 52(1974), 163-180.
- (43) M. Crandall and P. Rabinowitz. Perturbation of single eigenvalues and linearized stability. *Arch. Rat. Mech. Anal.*, 52(1974), 189-196.
- (44) M. Crandall and P. Rabinowitz. Some continuation and variational methods for nonlinear eigenvalue problems. *Bol. Soc. Mat. Esp.*, 56(1975), 157-184.
- (45) P. De Fazio and P. Di Napoli. Some monotone, uniqueness and stability results for a class of semilinear degenerate elliptic operators. *Boll. U.M.I.*, 3(1984), 239-251.
- (46) E.S. Noussair. Global solutions branching from positive eigenvalues. *Arch. Rat. Mech. Anal.*, 52(1974), 181-192.
- (47) E.W. Odeberg. On positive solutions of some pairs of differential equations. *Trans. Amer. Math. Soc.*, 249(1980), 729-743.
- (48) J.O. de Paiva-Pereira. Existence of boundary value problems of the Sturm-Liouville type. *Anais da Sociedade Brasileira de Matemática*, 19(1980), 209-225.
- (49) J.O. de Paiva-Pereira, Y.L. Linke and R.D. Nussbaum. A priori estimates and existence of positive solutions of semilinear elliptic equations. *J. Math. Pures Appl.*, 67(1987), 41-61.
- (50) D.G. de Figueiredo and R. Mitidieri. A variational principle for an elliptic system and applications to nonlinear problems. *N.B.C. Technical Report 255*, Rutgers, 1984.
- (51) R. Kacser. *Nonlinear Functional Analysis*. New York, Springer, 1985.
- (52) J.-L. Lions. *Nonlinear differential equations and free boundary problems*. London, Pitman, to appear.
- (53) J.-L. Lions and J. Peetre. On the existence of a free boundary for a class of reaction-diffusion systems. *SIAM J. Math. Anal.*, 15(1984), 610-625.
- (54) P.-C. Puel. *Mathematical aspects of semilinear and degenerate elliptic problems*. New York, Springer, Lecture Notes in Mathematics 80, 1979.
- (55) P.-C. Puel. Asymptotic series for equations of reaction and diffusion. *Bull. Amer. Math. Soc.*, 84(1978), 699-726.
- (56) P.-C. Puel and J.-M. Willem. The approach of solutions of nonlinear diffusion equations to equilibrium from reaction-diffusion theory. *Proc. Amer. Math. Soc.*, 84(1982), 282-286.
- (57) J. Serrin and R. Thompson. *Elliptic Partial Differential Equations of Second Order*. New York, Springer, 1973.
- (58) J. Serrin. Some problems in the theory of quasilinear elliptic equations. *Arch. Rat. Mech. Anal.*, 33(1970), 334-360.
- (59) J. Serrin and D. Philipp. The free boundary of a nonlinear elliptic equation. *Trans. Amer. Math. Soc.*, 282(1984), 107-122.
- (60) J. Serrin and R. Thompson. Elliptic Partial Differential Equations of Second Order. *Second edition*. New York, Springer, 1983.
- (61) J. Serrin. Some problems in the theory of quasilinear elliptic equations. *Arch. Rat. Mech. Anal.*, 33(1970), 334-360.
- (62) J. Serrin. Positive solutions and stability properties for solutions of reaction-diffusion systems with nonlinear boundary conditions. *Nonlinear Analysis, Methods and Applications*, 1(1976), 525-535.
- (63) J. Serrin. Positive solutions of reaction-diffusion systems with nonlinear boundary conditions. *Indiana Univ. Math. J.*, 27(1978), 165-177.
- (64) J. Serrin. Positive solutions of reaction-diffusion systems with nonlinear boundary conditions. To appear in Proc. Summer School, *Nonlinear Elliptic Equations in Mathematical Sciences*, V. Lakshmikanthan (ed.), New York, Academic Press, 1982, 525-535.
- (65) J. Serrin. Some free boundary problems for predator-prey systems with nonlinear diffusion. To appear in Proc. Summer School, *Nonlinear Elliptic Equations in Mathematical Sciences*, V. Lakshmikanthan (ed.), New York, Academic Press, 1982, 525-535.

- (67) G.-D. Joseph and T.S. Lundgren, Quasilinear Dirichlet problems driven by positive sources, *Arch. Ratl. Mech. Anal.*, **80** (1981), 241-269.
- (68) J.-L. Lions and P.-L. Lions, Remarks on some quasilinear elliptic problems, *Comm. Pure Appl. Math.*, **39** (1986), 545-569.
- (69) J.-P. Morel and L.-S. Nirenberg, Positive solutions of elliptic problems in Sobolev spaces, *J. Diff. Eq.*, **52**, 16 (1984), 197-225.
- (70) H.W. Hofer and E.R. Zehnder, Nonlinear problems suggested by nonlinear field equations, *Z. Math. Nach.*, **94** (1980), 135-156.
- (71) G.A. Kriegsmann, Stationary spectra patterns for a reaction-diffusion system with an unstable steady state, To appear.
- (72) G.A. Kriegsmann and E. Mihăilescu, Standing wave solutions for a system derived from the Fisher-Kolmogorov equation for nerve conduction, To appear in *SIAM J. Appl. Anal.*
- (73) H. Kressinski, Unstable periodic solutions in the theory of nonlinear heat transfer conduction, London, Longman Press, 1964.
- (74) N. Krasovskii, Positive solutions of operator equations, Gostekhizdat, Moscow, 1964.
- (75) A. Ladyzhenskaya and N. Ural'tseva, *Linear and quasilinear elliptic equations*, New York, Academic Press, 1968.
- (76) A.-S. Laetare and P.-J. Rabasse, On steady-state solutions of a system of reaction-diffusion equations from biology, *Nonlinear Anal.*, **6** (1992), 525-550.
- (77) J. Leger and J.-P. Béchade, Topology in aqueous emulsions, *Ann. Rev. Biophys. Biomol. Phys.*, **58** (1984), 45-76.
- (78) A. Leventhal, Reaction schemes for nonlinear elliptic systems related to ecology, *Arch. Ratl. Mech. Anal.*, **2** (1952), 277-295.
- (79) A. Leventhal and G. Cahn, Reaction and diffusion-adSORption behavior for perpendicularly transverse-elliptic boundary conditions with Dirichlet boundary data, *J. Diff. Eq.*, **35** (1980), 57-121.
- (80) F. Lin, A note on the existence of positive solutions of semilinear elliptic equations, *Adv. Math.*, **14** (1982), 487-497.
- (81) J.-B. McLeod, Qualitative methods in bifurcation theory, *Bifurc. Appl. Anal.*, **1** (1971), 42-61.
- (82) R.-H. Rand, Variational methods for nonlinear eigenvalue problems, in *Handbook of nonlinear partial differential equations*, G. Prokof'ev (ed.), Nauka, Moscow, 1974, 141-195.
- (83) P. Rabinowitz, Proof of positive solutions of nonlinear elliptic partial differential equations, *Inventiones Math.*, **23** (1974), 77-115.
- (84) P. Rabinowitz, On perturbation from symmetry I, *Diff. Eq.*, **14** (1971), 462-471.
- (85) P.-H. Rabinowitz, Positive solutions of nonlinear elliptic problems, in *Handbook of nonlinear partial differential equations*, G. Prokof'ev (ed.), Nauka, Moscow, 1974, 141-195.
- (86) P.-H. Rabinowitz, *Minimax methods in nonlinear analysis*, London, Peter W. Greylock, University Microfilms, 1973.
- (87) P.-H. Rabinowitz, Global minimization of functions of stationary solutions for a system of reaction-diffusion equations from biology, *Nonlinear Anal.*, **5** (1981), 497-508.
- (88) P.-H. Rabinowitz, Stability of bifurcation solutions by Leray-Schauder degree theory, *Arch. Ratl. Mech. Anal.*, **61** (1976), 154-164.
- (89) D.-H. Sattinger, Monotone methods in nonlinear elliptic and parabolic systems, *Inventiones Math.*, **31** (1976), 307-319.
- (90) D.-H. Sattinger, Stability of solutions of nonlinear equations, *J. Math. Anal. Appl.*, **19** (1977), 1-12.
- (91) D.-H. Sattinger, *Topics in nonlinear functional analysis*, Lecture Notes, New York, Gordon and Breach, 1976.
- (92) J.-P. Schröder, Variational and topological methods in nonlinear problems, *Arch. Ratl. Mech. Anal.*, **4** (1981), 267-295.
- (93) D.-H. Sattinger, The fixed point index for local condensing maps, *Annals Mat. Pura Appl.*, **89** (1971), 215-238.
- (94) G.-V. Teng, On nonlinear reaction-diffusion systems, *J. Math. Anal. Appl.*, **87** (1982), 145-159.
- (95) H.-K. Wong and A. Yosaki, Degenerate parabolic problems in population dynamics, To appear.
- (96) G. Zwick and A. Zivković, *Dynamical Models in Medicine*, Dordrecht, Kluwer, 1987.
- (97) M. Zhitova and G.-P. Winkler, *Nonlinear Problems in Differential Equations*, Princeton Univ. Press, Princeton, 1987.
- (98) P.-H. Rabinowitz, A note on nonlinear elliptic operator problems for a class of differential equations, *J. Diff. Eq.*, **9** (1971), 548-561.
- (99) P.-H. Rabinowitz, Some global results for nonlinear eigenvalue problems, *J. Funct. Anal.*, **7** (1971), 485-513.
- (100) P.-H. Rabinowitz, A general theorem for nonlinear eigenvalue problems and applications, *To Elliptic Operators in Nonlinear Functional Analysis*, H.H. Karcher (ed.), New York, Academic Press, 1973, 15-56.
- (101) P.-H. Rabinowitz, Some aspects of nonlinear eigenvalue problems, *Rocky Mountain J. Math.*, **3** (1973), 161-202.
- (102) P.-H. Rabinowitz, Proof of positive solutions of nonlinear elliptic partial differential equations, *Inventiones Math.*, **23** (1974), 77-115.

(101) R. S. Phillips and R. R. Coifman: Positive solutions of nonlinear elliptic eigenvalue problems. *J. Math. Mech.*, 19 (1970), 655-670.

(102) J. Serrin: *Local and global behavior of solutions of nonlinear elliptic equations*. New York: Springer-Verlag Notes in Mathematics 227, 1972, 228-269.

(103) G. H. Hardy, M. J. Riesz and G. Polya: *Some properties of fractional integrals, II*. Math. Z., 34 (1932), 403-439.

(104) L. Schwartz: *Équations aux dérivées partielles d'elliptique et un problème de Dirichlet*. Paris, 1963 (1971).

(105) L. Schwartz: *Équations aux dérivées partielles d'elliptiques et équation de la chaleur*. Paris, 1960.

(106) L. Schwartz: *Équations aux dérivées partielles d'elliptiques et équation de la chaleur*. New York: Academic Press, 1960.

(107) L. Schwartz: *Équations aux dérivées partielles d'elliptiques et équation de la chaleur*. Paris, 1960.

(108) L. Schwartz: *Équations aux dérivées partielles d'elliptiques et équation de la chaleur*. New York, 1960.

(109) L. Schwartz: *Équations aux dérivées partielles d'elliptiques et équation de la chaleur*. Paris, 1960.

(110) L. Schwartz: *Équations aux dérivées partielles d'elliptiques et équation de la chaleur*. Paris, 1960.

(111) R. E. L. Turner: *Nonlinear boundary value problems in nonlinear mechanics*. New York: Academic Press, 1966.

(112) R. E. L. Turner: *Nonlinear boundary value problems in nonlinear mechanics*. New York, 1966.

(113) R. E. L. Turner: *Nonlinear boundary value problems in nonlinear mechanics*. New York, 1966.

In this set of lectures, we shall investigate some nonstationary reaction-diffusion problems arising in chemical engineering. The range of phenomena that can occur is quite remarkable and we shall attempt to give a few problems that illustrate the rich variety of the field. Some important areas such as periodic oscillations and time propagation will not be touched upon at all. Among the important topics that deal with nonstationary reaction problems in chemical engineering, we mention here those of Aris (1975), Bodenstein and Lindner (1972), and Kirschhoff and Preuss (1947). I have also relied on a series of valuable review papers, including those of Aris (1974), Chaudhuri and Trivedi (1974), Goldstein, Gray and Wille (1974), and Rao (1977). Additional references on specific points will appear in the text.

1. THE BASIC EQUATIONS AND SOME ELEMENTARY APPROXIMATIONS

The basic equations governing chemical reactions are formulation of the conservation of mass and heat, the conservation of major physico-chemical quantities can be expressed in general as follows. Let Ω be a bounded domain in \mathbb{R}^n ($n = 1, 2, 3$) with boundary $\partial\Omega$. Let $\eta_{\Omega}(x)$ be the association or state vector in Ω at time t of the physical quantity under consideration (e.g. $\eta_{\Omega}(t)$). The problem is to find $\eta_{\Omega}(t)$ (that is the amount of the quantity being measured (perhaps enough to obtain pictures) per unit time in Ω at time t) (e.g. $\eta_{\Omega}(t)$ be the amount of the quantity flowing into Ω through $\partial\Omega$ per unit time).

Our basic conservation equation then takes the form

$$\frac{d\eta_{\Omega}}{dt} + \eta_{\Omega}' + \eta_{\Omega}'' = 0 \quad (1.1)$$

We shall assume that η_{Ω}' and η_{Ω}'' can be expressed as densities

$$\eta_{\Omega}(t) = \int q(x,t) dx + \eta_{\Omega}(0) - \int b(x,t) dx \quad (1.2)$$

We shall also assume that η_{Ω} can be written as a surface integral of

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