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AUTUMN COURSE ON APPLICATIONS OF ANALYSIS TO MECHANICS

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IDENTIFICATION OF SYSTEMS

(Part II)

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These are preliminary lecture notes intended for participants only.
Copies are available outside the Publications Office (T-floor) or from
room 112.

I Identification in an elliptic PDE: the continuous case.

We come back now to the identification problem of the § I.

Let $\Omega \subset \mathbb{R}^n$ be an open subset of $\mathbb{R}^n$ with boundary Γ , and let Γ_1, Γ_2 be a partition of Γ , and $\vec{n}$ be the normal to Γ pointing outwards of Ω .

For given functions a on Ω and g on Γ_2 , we consider the following elliptic PDE:

$$(P) \begin{cases} -\sum_{i=1}^n \frac{\partial}{\partial x_i} \left(a(x) \frac{\partial u}{\partial x_i} \right) = 0 & \text{on } \Omega \\ u = 0 & \text{on } \Gamma_1 \\ a(x) \frac{\partial u}{\partial n}(x) = g(x) & \text{on } \Gamma_2 \end{cases}$$

The function g being supposed known, to every function $a(x)$ we associate the error function $J(a)$ defined by

$$(4.1) \quad J(a) = \int_{\Omega} (u(x; a) - z(x))^2 dx$$

where $z(x)$ is a known function (representing some measure of the observed temperature distribution).

We shall first give a unity and existence theorem for the solution u of (P), so that the function $J(a)$ defined in (4.1) makes sense, and study the derivability of the application $a \rightarrow u$.

Then we shall calculate the derivative of J with respect to a and introduce the notion of adjoint state.

1) Study of the state equation

Multiply the first equation of (P) by a function $v(x)$ and integrate over Ω . Using the Green formula one gets:

$$\int_{\Omega} a(x) \sum_{i=1}^n \frac{\partial u}{\partial x_i} \frac{\partial v}{\partial x_i} - \int_{\Gamma} a(s) \frac{\partial u}{\partial n}(s) v(s) = 0 \quad \forall v$$

Taking $v = 0$ on Γ_1 we get, as $a(s) \frac{\partial u}{\partial n}(s) = g(s)$ on Γ_2 :

$$(4.2) \quad \int_{\Omega} a(x) \sum_{i=1}^n \frac{\partial u}{\partial x_i} \frac{\partial v}{\partial x_i} = \int_{\Gamma_2} g(s) v(s) \quad \forall v \text{ s.t. } v = 0 \text{ on } \Gamma_1$$

Suppose that

$$(4.3) \quad g \in L^2(\Gamma_2)$$

(4.4) Ω is bounded, Γ is "regular", and Γ_1 is of strictly positive $d\Gamma$ measure.
Define:

$$(4.5) \quad V = \{ u \in H^1(\Omega) \mid u = 0 \text{ on } \Gamma_1 \}$$

$$(4.6) \quad a(u, v) = \int_{\Omega} a(x) \sum_{i=1}^n \frac{\partial u}{\partial x_i} \frac{\partial v}{\partial x_i} \quad \forall u, v \in V$$

$$(4.7) \quad L(v) = \int_{\Gamma_2} g(s) v(s) \, ds \quad \forall v \in V$$

One deduces from (4.2) that the variational formulation of problem (P) is

$$(Q) \quad \boxed{\text{find } u \in V \text{ such that } a(u, v) = L(v) \quad \forall v \in V}$$

Define now:

$$(4.8) \quad \mathcal{A}_c = \{ a \in L^\infty(\Omega) \mid \exists \alpha > 0, \ a(x) \geq \alpha \text{ a.e. on } \Omega \}$$

which is obviously an open subset of $L^\infty(\Omega)$.

Thm 3 With the hypothesis and notation (4.2) to (4.8), the problem (Q) has a unique solution $u \in V$ for every $a \in \mathcal{A}_c$, and the mapping $a \rightarrow u$ is infinitely differentiable.

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Proof: As the trace application is a continuous mapping from $H^1(\Omega)$ in $L^2(\Gamma)$, V is a closed subspace of $H^1(\Omega)$, and hence a Hilbert space when equipped with the norm and scalar product of $H^1(\Omega)$.

The bilinear form $a(u, v)$ and the linear form $L(v)$ are obviously continuous on V .

Due to hypothesis (4.4) and Poincaré inequality, the bilinear form $a(u, v)$ is V -elliptic for every $a \in \mathcal{A}_c$, and the first part of the thm follows from Lax-Milgram theorem.

Let then $A(a)$ be the operator of $\mathcal{L}(V, V')$ associated to the bilinear form $a(u, v)$ (or to the function $a(x)$) by:

$$(4.9) \quad (A(a)u, v) = a(u, v) \quad \forall u, v \in V$$

So the equation (Q) may be rewritten

$$(4.10) \quad A(a)u = L$$

or equivalently, as $A(a)$ is invertible (Lax-Milgram):

$$(4.11) \quad u = [A(a)]^{-1} L$$

Hence the application $a \rightarrow u$ from $\mathcal{A}_c$ in V is the composition of two applications:

- the application $a \rightarrow A(a)$ which is obviously linear continuous from $\mathcal{A}_c$ in $\text{Isom}(V, V')$ (1)

- the application $A \rightarrow A^{-1}$ from $\text{Isom}(V, V')$ in $\text{Isom}(V', V)$ which is known to be $\mathcal{C}^\infty$

So the application $a \rightarrow u$ is $\mathcal{C}^\infty$ from $\mathcal{A}_c$ in V , which ends the proof of thm 3. ■

2) Adjoint state and derivability of J .

Let us denote by $(,)$ and $||$ the scalar product and norm in $L^2(\Omega)$. Then the error function J defined in (4.1) may be rewritten:

$$(4.12) \quad J(a) = ||u(a) - \bar{z}||^2$$

where $u(a)$ is the solution of (Q) for the value a of the unknown parameter.

It follows from Thm 3 that J is a C^∞ function of a , and we want now to calculate its first derivative.

The most convenient way to do this is to consider J as a function of two variables u and a related by the equation (Q). So we are lead to introduce a Lagrangian function depending on u, a and on a Lagrange multiplier p for the equation (Q):

$$(4.13) \quad \begin{cases} \forall (u, a) \in V \times \mathcal{A}, \quad \forall p \in V & \text{define:} \\ \mathcal{L}(u, a; p) = ||u - \bar{z}||^2 + a(u, p) - L(p) \end{cases}$$

Then the error criterion is given (for any $p \in V$!) by:

$$(4.14) \quad J(a) = \mathcal{L}(u(a), a; p)$$

As we know from thm 3 that the mapping $a \rightarrow u(a)$ is derivable, we may differentiate (4.14) with respect to a :

$$(4.15) \quad \delta J = \frac{\partial \mathcal{L}}{\partial u}(u(a), a; p) \delta u + \frac{\partial \mathcal{L}}{\partial a}(u(a), a; p) \delta a$$

The equation (4.15) is true for any $p \in V$. so if we choose p such that:

$$(4.16) \quad \frac{\partial \mathcal{L}}{\partial u}(u(a), a; p) \cdot v = 0 \quad \forall v \in V$$

we get from (4.15):

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$$(4.17) \quad \delta J = \frac{\partial \mathcal{L}}{\partial a} (u(a), a; p) \delta a$$

which is the desired first derivative of J with respect to a .

Let us now interpret (4.16) (4.17).

Using (4.13), (4.16) becomes

$$(\tilde{Q}) \quad \boxed{\text{find } p \in V \text{ s.t. } a(u, p) = -2(u - \bar{z}, v) \quad \forall v \in V}$$

which has effectively a unique solution $p \in V$ when $a \in \mathcal{A}_c$ (the bilinear form $a(u, v)$ is still V -elliptic), and (4.17) gives then:

$$(4.18) \quad \delta J = \delta a(u, p) = \int_{\Omega} \delta a(x) \sum_{i=1}^n \frac{\partial u}{\partial x_i} \frac{\partial p}{\partial x_i} dx$$

The function p is called the adjoint state, and the equation $(\tilde{Q})$ is the adjoint state equation. It is easy to check that the adjoint state p is a weak solution of:

$$(\tilde{P}) \quad \begin{cases} - \sum_{i=1}^n \frac{\partial}{\partial x_i} \left(a(x) \frac{\partial p}{\partial x_i} \right) = -2(u - \bar{z}) & \text{in } \Omega \\ p = 0 & \text{on } T_1 \\ a(x) \frac{\partial p}{\partial n}(x) = 0 & \text{on } T_2 \end{cases}$$

which is to be compared with the (dual) state equation (P). So we have proved the

Thm 4: Under the hypothesis of thm 3, the error function J is C^∞ with respect to a on $\mathcal{A}_c$. Its first derivative is given by (4.18), where $u \in V$ is the (dual) state solution of (Q), and where $p \in V$ is the adjoint state solution of $(\tilde{Q})$.

Remark 6: As $a \in L^\infty(\Omega)$, $J'(a)$ is in $[L^\infty(\Omega)]' \supset L^1(\Omega)$, and (2.18) shows that $J'(a)$ is in fact in $L^1(\Omega)$: define

$$(4.19) \quad \frac{\partial J}{\partial a}(x) = \sum_{i=1}^n \frac{\partial u}{\partial x_i}(x) \frac{\partial p}{\partial x_i}(x) \quad \forall x \in \Omega$$

so that (4.18) rewrites:

$$(4.20) \quad \delta J = \int_{\Omega} \frac{\partial J}{\partial a}(x) \delta a(x) \, dx$$

The function $\frac{\partial J}{\partial a}$, which belongs to $L^1(\Omega)$, is termed the partial functional derivative of J with respect to a . Its value $\frac{\partial J}{\partial a}(x)$ at a point x of Ω may be seen as the partial derivative of J with respect to the value $a(x)$ of a at x .

If the functions a and $\frac{\partial J}{\partial a}$ would belong to $L^2(\Omega)$, then $\frac{\partial J}{\partial a}$ would be exactly the gradient of J with respect to a , and we could design a gradient method by defining a sequence $a^0, a^1, \dots$ etc. by

$$(4.21) \quad a^{n+1}(x) = a^n(x) - \rho \frac{\partial J}{\partial a}(a^n)(x) \quad \forall x \in \Omega$$

As the function $\frac{\partial J}{\partial a}(x)$ is in fact only in $L^1(\Omega)$ when $a(x)$ is in $L^\infty(\Omega)$, a method such as (4.21) does not make sense (a^{n+1} does not belong to L^∞ !), and the gradient method sketched in § II and III has to be applied on the discretized problem (see § IV) ■

Remark 7: The calculation of the partial functional derivative $\frac{\partial J}{\partial a}(x)$ requires only the knowledge of u and p , i.e. the resolution of two partial differential equations, though the parameter space is infinite dimensional! ■

Exercise 1 Prove thm 4 for the state equation (P) - a(4) - but with a boundary observation on Γ_2 , i.e. replace (4.2) by

$$(4.22) \quad J(a) = \int_{\Gamma_2} (u(s; a) - z(s))^2 ds$$

where z is a given function of $L^2(\Gamma_2)$

Exercise 2 Prove thm 4 for the state equation:

$$(P') \begin{cases} - \sum_{i,j=1}^n \frac{\partial}{\partial x_i} \left(a_{ij}(x) \frac{\partial u}{\partial x_j} \right) = 0 & \text{on } \Omega \\ u = 0 & \text{on } \Gamma_1 \\ a(s) \frac{\partial u}{\partial n}(s) = \tau(s) (u_e(s) - u(s)) & \text{on } \Gamma_2 \end{cases}$$

where $a \in L^\infty(\Omega)$ and $u_e \in L^2(\Gamma_2)$ are known, with $a_{ij}(x) \geq \alpha > 0$, and where $\tau(s)$ is an unknown positive function; the observation being on the boundary Γ_2 as in Ex 1.

Exercise 3 Prove thm 4 for the state equation:

$$(P'') \begin{cases} - \sum_{i,j=1}^n \frac{\partial}{\partial x_i} \left(a_{ij} \frac{\partial u}{\partial x_j} \right) = f & \text{in } \Omega \\ u = 0 & \text{on } \Gamma \end{cases}$$

where $f \in L^2(\Omega)$ is known, and where $a_{ij} \in L^\infty(\Omega)$, $i, j = 1, \dots, n$ are unknown functions, such that for almost every $x \in \Omega$, the matrix of coefficients a_{ij} is positive definite; The observation being taken at your convenience.

Exercise 4 Prove thm 4 for the state equations (P) and for point-measurements. A point-measurement of u at point x_i may be seen as the mean of the mean value of u on a small domain Ω_i surrounding x_i :

$$(4.23) \quad J(a) = \sum_{i=1}^K \left(\frac{1}{\text{meas } \Omega_i} \int_{\Omega_i} u(x; a) dx - z_i \right)^2$$

where K is the number of point-measurements and $z \in \mathbb{R}^K$ is a

Ⓐ Identification in an elliptic PDE: the discretized case.

As it is impossible to "compute" the function $u(x)$ solution of (Q), we have to replace it by a function $u_h(x)$ depending on a finite number of degree of liberty, so that we may determine it on a computer. We will show in this § how the use of a variational formulation as (Q) makes it possible to design easily numerical approximation schemes.

1) A glossary of approximation theory for variational problems

As suggested by the example of § Ⓓ, let

$$(5.1) \quad V = \text{Hilbert Space, with norm } \|\cdot\| \text{ and scalar product } (\cdot, \cdot)$$

$$(5.2) \quad a: V \times V \rightarrow \mathbb{R} \text{ a bilinear continuous form}$$

$$(5.3) \quad L: V \rightarrow \mathbb{R} \text{ a linear continuous form.}$$

We know that, under the hypothesis

$$(5.4) \quad \exists \alpha > 0, \quad a(v, v) \geq \alpha \|v\|^2 \quad (\text{ie } a \text{ is } V\text{-elliptic})$$

the problem

$$(Q) \quad \text{find } u \in V \text{ s.t. } a(u, v) = L(v) \quad \forall v \in V$$

has a unique solution.

Let h be an index (which shall tend to zero), which, in practical applications, shall be related to the "mesh size" of the approximation.

To every such h , we associate a finite dimensional subspace V_h of V , ie:

$$(5.5) \quad \forall h, \quad \dim V_h = I(h) < +\infty$$

and we define the approximated problem (Q_h) by:

$$(Q_h) \quad \text{find } u_h \in V_h \text{ s.t. } a(u_h, v_h) = L(v_h) \quad \forall v_h \in V_h$$

We have first the:

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Thm 5 With the hypothesis and notation (5.1) to (5.5), the problem
 (Q_h) has a unique solution u_h , related to the solution
 u of (6) by :

$$(5.6) \quad \|u_h - u\| \leq \frac{M}{\alpha} \inf_{v_h \in V_h} \|v_h - u\|$$

where M and α are respectively the continuity and V -ellipticity constants of the bilinear form a .

Proof. The uniqueness and existence of u_h results obviously from the application of the Lax-Milgram theorem to the problem (Q_h) .

Taking $v = v_h$ in (6) and subtracting from (Q_h) gives:

$$a(u_h - u, v_h) = 0 \quad \forall v_h \in V_h$$

so that

$$a(u_h - u, u_h - u) = a(u_h - u, v_h - u) \quad \forall v_h \in V$$

Using the continuity and V -ellipticity of a gives

$$\alpha \|u_h - u\|^2 \leq M \|u_h - u\| \|v_h - u\| \quad \forall v_h \in V$$

which gives (5.6). ■

As we want that the approximate solution u_h tends toward u when $h \rightarrow 0$, we have to make some hypothesis on the choice of the approximation spaces V_h . A reasonable hypothesis seems to ask that there exists a family of linear applications I_h from V in V_h such that, for every $v \in V$ (as we don't know u !), $I_h v \rightarrow v$ in V as $h \rightarrow 0$. In fact it is sufficient that this property holds on a dense subset $\mathcal{D}$ of V , as stated in the

Thm 6 Under the hypothesis of thm 5 and if there exists a dense subspace $\mathcal{Q}$ of V and a family of linear continuous applications $\mathcal{I}_h$ of $\mathcal{Q}$ in V_h such that

$$(5.7) \quad \forall v \in \mathcal{Q} \quad \mathcal{I}_h v \rightarrow v \text{ in } V \text{ as } h \rightarrow 0$$

then

$$(5.8) \quad u_h \rightarrow u \text{ in } V \text{ as } h \rightarrow 0$$

i.e. the approximated solution u_h tends toward the exact solution u as the "mesh size" h tends to zero.

Proof: Let $\varepsilon > 0$ be given. As $\mathcal{Q}$ is dense in V , there exists a $v \in \mathcal{Q}$ such that

$$\|u - v\| \leq \frac{\varepsilon}{2H}$$

As $\mathcal{I}_h v \rightarrow v$, $\exists h_0$ such that

$$h \leq h_0 \Rightarrow \|\mathcal{I}_h v - v\| \leq \frac{\varepsilon}{2H}$$

so that

$$\|u - \mathcal{I}_h v\| \leq \frac{\varepsilon}{H}$$

taking then $v_h = \mathcal{I}_h v$ in (5.6) gives:

$$\|u_h - u\| \leq \varepsilon \text{ when } h \leq h_0$$

Remark 8 The definition of the approximated problem (4) requires only the choice of the family of spaces V_h . The choice of $\mathcal{Q}$ and $\mathcal{I}_h$ is "necessary" only for the proof of the convergence of the approximation.

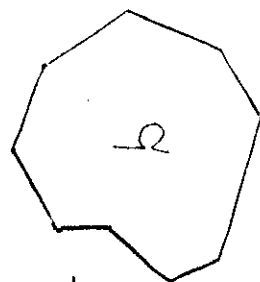
Remark 9: The thm 6 proves only that $u_h \rightarrow u$, but does not give any estimation for the error made when replacing u by u_h . It follows from formula (5.6) such estimation of error always requires some a-priori knowledge on the exact solution u , as u appears in the right hand side of (5.6)!

2) Finite elements approximation of the space $H^1(\Omega)$

We shall construct, in that paragraph, a family of subspaces V_h of $H^1(\Omega)$ for the case:

$$(5.9) \quad \Omega \subset \mathbb{R}^2, \quad \Omega \text{ is polygonal}$$

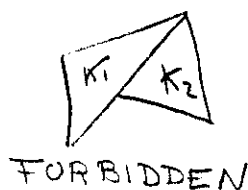
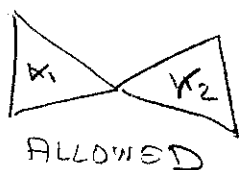
using the technique of finite elements.



To every $h > 0$ we make correspond a finite family $\mathcal{B}_h$ of closed triangles K such that:

$$(5.10) \quad \bar{\Omega} = \bigcup_{K \in \mathcal{B}_h} K$$

$$(5.11) \quad \left\{ \begin{array}{l} \text{given two triangles } K_1 \text{ and } K_2 \text{ of } \mathcal{B}_h, \text{ the only} \\ \text{possible situations are:} \\ K_1 \cap K_2 = \emptyset \\ K_1 \cap K_2 = \text{1 common side} \\ K_1 \cap K_2 = \text{1 common vertex} \end{array} \right.$$



$$(5.12) \quad \text{diam } K \leq h \quad \forall K \in \mathcal{B}_h$$

The subspaces V_h of $H^1(\Omega)$ are then defined by: ⁽¹⁾

$$(5.13) \quad V_h = \left\{ v_h \in C^0(\bar{\Omega}) \mid v_h|_K \in \mathbb{P}^1 \quad \forall K \in \mathcal{B}_h \right\}$$

where $\mathbb{P}^1$ denotes the space of polynomials of degree ≤ 1 in two variables.

Define (cf fig 7)

$$(5.14) \quad \Omega_h = \{ \text{all vertex of } K \text{ with } K \in \mathcal{B}_h \}, \quad I(h) = \text{card } \Omega_h$$

Then obviously we have the

Thm 7: The subspaces V_h of $H^1(\Omega)$ defined by (5.13) are finite-dimensional, and precisely:

$$(5.15) \quad \dim V_h = I(h) < +\infty \quad \forall h > 0$$

A function $v_h \in V_h$ is uniquely determined by the data of its values on the vertex in Ω_h (those values are called "the degree of freedom of the space V_h ")

The corresponding basis of V_h is the collection of functions $\{w_M, M \in \Omega_h\}$ defined by:

$$(5.16) \quad w_M(P) = \delta_{MP} = \begin{cases} 0 & \text{if } M \neq P \\ 1 & \text{if } M = P \end{cases} \quad \forall P \in \Omega_h$$

It's easy to check that

$$(5.17) \quad \text{Supp } w_M = \bigcup_{M \text{ is a vertex of } K} K$$

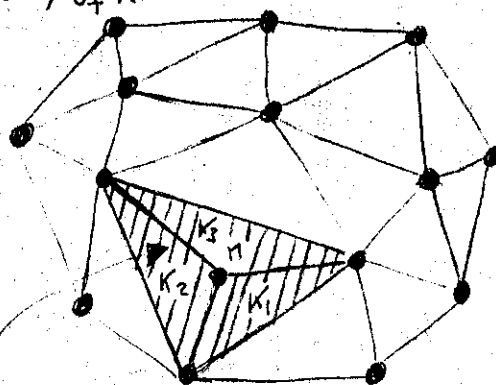
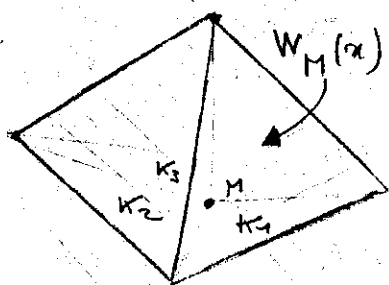


Fig 7 the basis function w_M , its support, $\Omega = \bigcup \text{triangles}$, $\Omega_h = \{\bullet\}$
With that choice of basis functions we have:

$$(5.18) \quad \forall v_h \in V_h, \quad v_h(x) = \sum_{M \in \Omega_h} v_h(M) w_M(x)$$

Let us now turn to the convergence of this approximation of $H^1(\Omega)$: we have to construct a dense subspace $\mathcal{Q}$ of V and a family $\mathcal{Q}_h$ of restriction of $\mathcal{Q}$ to V_h .

Define the inradius diameter ρ of any triangle K by:

$$(5.19) \quad \rho(K) = \sup_{\rho > 0 \text{ s.t. } \exists \text{ disk of diameter } \rho \text{ included in } K} \rho$$

$$(5.20) \quad \rho_h = \min_{K \in \mathcal{B}_h} \rho(K)$$

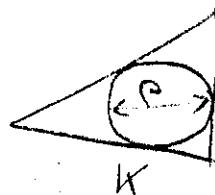
Then if we choose

$$(5.21) \quad \mathcal{V} = \{v \mid \exists w \in \mathcal{C}^2(\mathbb{R}^2), \text{ s.t. } v = w|_{\sqrt{\epsilon}}\}$$

and:

$$(5.22) \quad \mathcal{I}_h: \forall v \in \mathcal{C}^2(\sqrt{\epsilon}) \longrightarrow v_h = \mathcal{I}_h v = \sum_{M \in \mathcal{S}_h} v(M) w_M$$

we have the



Theorem 8: With the hypothesis and notation (5.9) to (5.14) and

(5.19) to (5.22), for every $v \in \mathcal{V}$, the function $v_h = \mathcal{I}_h v$ is related to v by the following majorations:

$$(5.23) \quad \forall x \in \Omega, \quad |v_h(x) - v(x)| \leq \frac{M}{2} h^2$$

$$(5.24) \quad \forall K \in \mathcal{B}_h, \forall x \in K, \quad \|v_h'(x) - v'(x)\| \leq \frac{M}{2} h^2 \frac{3C}{\rho}$$

where C is a constant independent of h , and where:

$$(5.25) \quad M = \max_{x \in \overline{\Omega}} \|v''(x)\|$$

Hence:

$$(5.26) \quad \|v_h - v\|_{H^1(\Omega)} \leq \frac{1}{2} M \max_{\Omega} h \sqrt{h + 3C \frac{h}{\rho}}$$

Let us first give some lemmas:

Lemma 1: $\forall v \in \mathcal{V}$ and $v_h = \mathcal{I}_h v$ defined by (5.22) we have:

$$(5.27) \quad v_h(x) = v(x) + \sum_{M \in \mathcal{S}_h} \frac{1}{2} v''(\tilde{x}_M) (M-x)^2 w_M(x) \quad \forall x \in \Omega$$

$$(5.28) \quad v_h'(x) = v'(x) + \sum_{M \in \mathcal{S}_h} \frac{1}{2} v''(\tilde{x}_M) (M-x)^2 w_M'(x) \quad \begin{matrix} \forall K \in \mathcal{B}_h \\ \forall x \in K \end{matrix}$$

where $\tilde{x}_M \in [x, M] \quad \forall M \in \mathcal{S}_h$.

Proof (Exercise!) For every $M \in \mathcal{S}_h$, substitute the second order development

$$(5.29) \quad v(M) = v(x) + v'(x)(M-x) + \frac{1}{2} v''(\tilde{x}_M) (M-x)^2$$

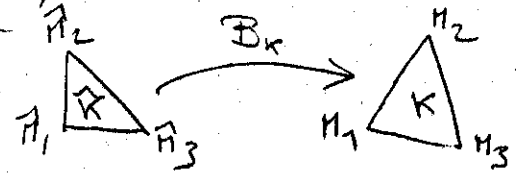
in the definition (5.22) of $v_h(x)$. Take into account that $\sum_{M \in \mathcal{S}_h} w_M(x) = 1$.

$v \in P^\perp$ gives (5.27). Similarly, substituting (5.29) in (5.30) gives (5.28).

$$v'_h(x) = \sum_{M \in \mathcal{S}_h} v(M) w'_M(x)$$

Lemma 2: Let $\hat{K}$ be a reference triangle in a reference plane with coordinates $\hat{x} = (\hat{x}_1, \hat{x}_2)$. For every $K \in \mathcal{T}_h$, we denote by B_K the (uniquely defined) linear application from $\mathbb{R}^2$ in $\mathbb{R}^2$ such that

(5.31)
$$\begin{cases} B_K \hat{M}_i = M_i & i=1,2,3 \\ \text{where } \hat{M}_i \text{ and } M_i \text{ are the vertices of } \hat{K} \text{ and } K \end{cases}$$



Then

$$(5.32) \quad \|B_K\| \leq \frac{h}{\hat{\rho}}, \quad \|B_K^{-1}\| \leq \frac{\hat{h}}{\rho}$$

where $\hat{h}$ and $\hat{\rho}$ are associated to $\hat{K}$ as h and ρ to K .

Proof:
$$\|B_K\| = \sup_{\hat{z} \neq 0} \frac{\|B_K \hat{z}\|}{\|\hat{z}\|} = \sup_{\|\hat{z}\| = \hat{\rho}} \frac{\|B_K \hat{z}\|}{\hat{\rho}}$$

As $\|\hat{z}\| = \hat{\rho}$, $\exists \hat{x}_1$ and $\hat{x}_2 \in \hat{K}$ s.t. $\hat{z} = \hat{x}_1 - \hat{x}_2$. Hence:

$$\begin{aligned} \|B_K \hat{z}\| &= \|B_K \hat{x}_1 - B_K \hat{x}_2\| \\ &\leq h \quad \text{as both } B_K \hat{x}_1 \text{ and } B_K \hat{x}_2 \text{ belong to } K. \end{aligned}$$

Proof of Thm 8 Using Lemma 1 and (5.25) gives

$$|v_h(u) - \pi(u)| \leq \frac{M}{2} \sum_{M \in \mathcal{S}_h} \|M - x\|^2 |w_M(x)| \quad \forall x \in \mathcal{E}$$

$$\|v_h(u) - v(u)\| \leq \frac{M}{2} \sum_{M \in \mathcal{S}_h} \|M - x\|^2 \|w'_M(x)\| \quad \forall K \in \mathcal{T}_h, \forall x \in K$$

For a given $M \in \mathcal{S}_h$, either $x \in \text{supp } w_M$, and then

$$|v_h(x) - v(x)| \leq \frac{M}{2} h^2 \sum_{T \in \mathcal{T}_h} |w_T(x)|$$

forall x in R

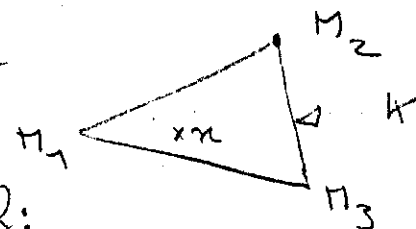
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$$(5.33) \quad \|v_h'(x) - v'(x)\| \leq \frac{M}{2} h^2 \sum_{T \in \mathcal{T}_h} \|w_T'(x)\| \quad \forall x \in \mathcal{E}_h, \forall x \in \mathcal{K}$$

It's easy to check that $\sum_{T \in \mathcal{T}_h} |w_T(x)| = 1$ which gives (5.23).

We have now to get a majoration for $\sum_{T \in \mathcal{T}_h} \|w_T'(x)\|$.

As $x \in \mathcal{K}$, This summation contains only three non-zero terms, corresponding to the three vertex π_1, π_2, π_3 of $\mathcal{K}$.



Then we have, with the notations of Lemma 2:

$$\begin{cases} w_{\pi_i}(x) = \hat{w}_i(B_K^{-1}x) & i=1,2,3 \end{cases}$$

where $\hat{w}_i$ is the linear function of $(\hat{\pi}_1, \hat{\pi}_2)$ s.t. $\hat{w}_i(\hat{\pi}_j) = \delta_{ij}$

so that:

$$(5.34) \quad w_{\pi_i}'(x) = \hat{w}_i'(\hat{x}) B_K^{-1} \quad \text{with} \quad \hat{x} = B_K^{-1}x \in \hat{\mathcal{K}}$$

and, using (5.32)

$$(5.35) \quad \|w_{\pi_i}'(x)\| \leq C/\rho$$

with

$$(5.36) \quad C = \hat{h} \max_{i=1,2,3} \|\hat{w}_i'\| \quad \text{is independent of } h$$

independent of $\hat{x}$ as $\hat{w}_i$ is linear!

Then using (5.35) we get:

$$\sum_{T \in \mathcal{T}_h} \|w_T'(x)\| \leq 3C/\rho \quad \text{with } C = \text{const indep of } h$$

which gives (5.24)

The majoration (5.26) is a simple corollary of (5.23)(5.24). This ends the proof of thm 8. $\blacksquare$

The thm 8 shows that, if we want that the approximation V_h is converging, it is not sufficient that the size h of all triangles $\mathcal{T}$ tends to zero: we see from (5.26) that the ratio h/ρ of exterior to interior diameter of all triangles has to be bounded.

Corollary of thm 8: With the hypothesis of thm 8 and

(5.37) $\frac{h}{\rho}$ bounded as $h \rightarrow 0$

or equivalently

(5.38) the smallest angle of the triangles $\mathcal{T}$ of $\mathcal{T}_h$ is bounded below as $h \rightarrow 0$

then the approximation of $H^1(\Omega)$ by the family of spaces V_h defined in (5.13) is convergent when $h \rightarrow 0$.

Proof: It results from (5.26) and thm 6, and from the fact that the space $\mathcal{V}$ defined by (5.21) is dense in $H^1(\Omega)$ (see (5.22)). ■